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# Shortlisted 

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Shortlisted Problems with Solutions

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## Contributing Countries

Austria, Australia, Belgium, Bulgaria, Canada, Croatia, Czech Republic, Estonia, Finland, Greece, India, Indonesia, Iran, Japan, Korea (North), Korea (South), Lithuania, Luxembourg, Mexico, Moldova, Netherlands, New Zealand, Poland, Romania, Russia, Serbia, South Africa, Sweden, Thailand, Taiwan, Turkey, Ukraine, United Kingdom, United States of America

## Problem Selection Committee

Ha Huy Khoai
Ilya Bogdanov
Tran Nam Dung
Le Tuan Hoa
Géza Kós

## Algebra

A1. Given a sequence $a_{1}, a_{2}, \ldots, a_{n}$ of real numbers. For each $i(1 \leq i \leq n)$ define

$$
d_{i}=\max \left\{a_{j}: 1 \leq j \leq i\right\}-\min \left\{a_{j}: i \leq j \leq n\right\}
$$

and let

$$
d=\max \left\{d_{i}: 1 \leq i \leq n\right\} .
$$

(a) Prove that for arbitrary real numbers $x_{1} \leq x_{2} \leq \ldots \leq x_{n}$,

$$
\begin{equation*}
\max \left\{\left|x_{i}-a_{i}\right|: 1 \leq i \leq n\right\} \geq \frac{d}{2} \tag{1}
\end{equation*}
$$

(b) Show that there exists a sequence $x_{1} \leq x_{2} \leq \ldots \leq x_{n}$ of real numbers such that we have equality in (1).
(New Zealand)
Solution 1. (a) Let $1 \leq p \leq q \leq r \leq n$ be indices for which

$$
d=d_{q}, \quad a_{p}=\max \left\{a_{j}: 1 \leq j \leq q\right\}, \quad a_{r}=\min \left\{a_{j}: q \leq j \leq n\right\}
$$

and thus $d=a_{p}-a_{r}$. (These indices are not necessarily unique.)


For arbitrary real numbers $x_{1} \leq x_{2} \leq \ldots \leq x_{n}$, consider just the two quantities $\left|x_{p}-a_{p}\right|$ and $\left|x_{r}-a_{r}\right|$. Since

$$
\left(a_{p}-x_{p}\right)+\left(x_{r}-a_{r}\right)=\left(a_{p}-a_{r}\right)+\left(x_{r}-x_{p}\right) \geq a_{p}-a_{r}=d,
$$

we have either $a_{p}-x_{p} \geq \frac{d}{2}$ or $x_{r}-a_{r} \geq \frac{d}{2}$. Hence,

$$
\max \left\{\left|x_{i}-a_{i}\right|: 1 \leq i \leq n\right\} \geq \max \left\{\left|x_{p}-a_{p}\right|,\left|x_{r}-a_{r}\right|\right\} \geq \max \left\{a_{p}-x_{p}, x_{r}-a_{r}\right\} \geq \frac{d}{2}
$$

(b) Define the sequence $\left(x_{k}\right)$ as

$$
x_{1}=a_{1}-\frac{d}{2}, \quad x_{k}=\max \left\{x_{k-1}, a_{k}-\frac{d}{2}\right\} \quad \text { for } 2 \leq k \leq n
$$

We show that we have equality in (1) for this sequence.
By the definition, sequence $\left(x_{k}\right)$ is non-decreasing and $x_{k}-a_{k} \geq-\frac{d}{2}$ for all $1 \leq k \leq n$. Next we prove that

$$
\begin{equation*}
x_{k}-a_{k} \leq \frac{d}{2} \quad \text { for all } 1 \leq k \leq n \tag{2}
\end{equation*}
$$

Consider an arbitrary index $1 \leq k \leq n$. Let $\ell \leq k$ be the smallest index such that $x_{k}=x_{\ell}$. We have either $\ell=1$, or $\ell \geq 2$ and $x_{\ell}>x_{\ell-1}$. In both cases,

$$
\begin{equation*}
x_{k}=x_{\ell}=a_{\ell}-\frac{d}{2} . \tag{3}
\end{equation*}
$$

Since

$$
a_{\ell}-a_{k} \leq \max \left\{a_{j}: 1 \leq j \leq k\right\}-\min \left\{a_{j}: k \leq j \leq n\right\}=d_{k} \leq d,
$$

equality (3) implies

$$
x_{k}-a_{k}=a_{\ell}-a_{k}-\frac{d}{2} \leq d-\frac{d}{2}=\frac{d}{2} .
$$

We obtained that $-\frac{d}{2} \leq x_{k}-a_{k} \leq \frac{d}{2}$ for all $1 \leq k \leq n$, so

$$
\max \left\{\left|x_{i}-a_{i}\right|: 1 \leq i \leq n\right\} \leq \frac{d}{2}
$$

We have equality because $\left|x_{1}-a_{1}\right|=\frac{d}{2}$.
Solution 2. We present another construction of a sequence ( $x_{i}$ ) for part (b).
For each $1 \leq i \leq n$, let

$$
M_{i}=\max \left\{a_{j}: 1 \leq j \leq i\right\} \quad \text { and } \quad m_{i}=\min \left\{a_{j}: i \leq j \leq n\right\}
$$

For all $1 \leq i<n$, we have

$$
M_{i}=\max \left\{a_{1}, \ldots, a_{i}\right\} \leq \max \left\{a_{1}, \ldots, a_{i}, a_{i+1}\right\}=M_{i+1}
$$

and

$$
m_{i}=\min \left\{a_{i}, a_{i+1}, \ldots, a_{n}\right\} \leq \min \left\{a_{i+1}, \ldots, a_{n}\right\}=m_{i+1} .
$$

Therefore sequences $\left(M_{i}\right)$ and $\left(m_{i}\right)$ are non-decreasing. Moreover, since $a_{i}$ is listed in both definitions,

$$
m_{i} \leq a_{i} \leq M_{i}
$$

To achieve equality in (1), set

$$
x_{i}=\frac{M_{i}+m_{i}}{2} .
$$

Since sequences $\left(M_{i}\right)$ and $\left(m_{i}\right)$ are non-decreasing, this sequence is non-decreasing as well.

From $d_{i}=M_{i}-m_{i}$ we obtain that

$$
-\frac{d_{i}}{2}=\frac{m_{i}-M_{i}}{2}=x_{i}-M_{i} \leq x_{i}-a_{i} \leq x_{i}-m_{i}=\frac{M_{i}-m_{i}}{2}=\frac{d_{i}}{2} .
$$

Therefore

$$
\max \left\{\left|x_{i}-a_{i}\right|: 1 \leq i \leq n\right\} \leq \max \left\{\frac{d_{i}}{2}: 1 \leq i \leq n\right\}=\frac{d}{2}
$$

Since the opposite inequality has been proved in part (a), we must have equality.

A2. Consider those functions $f: \mathbb{N} \rightarrow \mathbb{N}$ which satisfy the condition

$$
\begin{equation*}
f(m+n) \geq f(m)+f(f(n))-1 \tag{1}
\end{equation*}
$$

for all $m, n \in \mathbb{N}$. Find all possible values of $f(2007)$.
( $\mathbb{N}$ denotes the set of all positive integers.)
(Bulgaria)
Answer. 1, 2, ..., 2008.
Solution. Suppose that a function $f: \mathbb{N} \rightarrow \mathbb{N}$ satisfies (1). For arbitrary positive integers $m>n$, by (1) we have

$$
f(m)=f(n+(m-n)) \geq f(n)+f(f(m-n))-1 \geq f(n),
$$

so $f$ is nondecreasing.
Function $f \equiv 1$ is an obvious solution. To find other solutions, assume that $f \not \equiv 1$ and take the smallest $a \in \mathbb{N}$ such that $f(a)>1$. Then $f(b) \geq f(a)>1$ for all integer $b \geq a$.

Suppose that $f(n)>n$ for some $n \in \mathbb{N}$. Then we have

$$
f(f(n))=f((f(n)-n)+n) \geq f(f(n)-n)+f(f(n))-1
$$

so $f(f(n)-n) \leq 1$ and hence $f(n)-n<a$. Then there exists a maximal value of the expression $f(n)-n$; denote this value by $c$, and let $f(k)-k=c \geq 1$. Applying the monotonicity together with (1), we get

$$
\begin{aligned}
2 k+c \geq f(2 k)=f(k+k) & \geq f(k)+f(f(k))-1 \\
& \geq f(k)+f(k)-1=2(k+c)-1=2 k+(2 c-1)
\end{aligned}
$$

hence $c \leq 1$ and $f(n) \leq n+1$ for all $n \in \mathbb{N}$. In particular, $f(2007) \leq 2008$.
Now we present a family of examples showing that all values from 1 to 2008 can be realized. Let

$$
f_{j}(n)=\max \{1, n+j-2007\} \quad \text { for } j=1,2, \ldots, 2007 ; \quad f_{2008}(n)= \begin{cases}n, & 2007 \nmid n \\ n+1, & 2007 \mid n\end{cases}
$$

We show that these functions satisfy the condition (1) and clearly $f_{j}(2007)=j$.
To check the condition (1) for the function $f_{j}(j \leq 2007)$, note first that $f_{j}$ is nondecreasing and $f_{j}(n) \leq n$, hence $f_{j}\left(f_{j}(n)\right) \leq f_{j}(n) \leq n$ for all $n \in \mathbb{N}$. Now, if $f_{j}(m)=1$, then the inequality (1) is clear since $f_{j}(m+n) \geq f_{j}(n) \geq f_{j}\left(f_{j}(n)\right)=f_{j}(m)+f_{j}\left(f_{j}(n)\right)-1$. Otherwise,

$$
f_{j}(m)+f_{j}\left(f_{j}(n)\right)-1 \leq(m+j-2007)+n=(m+n)+j-2007=f_{j}(m+n) .
$$

In the case $j=2008$, clearly $n+1 \geq f_{2008}(n) \geq n$ for all $n \in \mathbb{N}$; moreover, $n+1 \geq$ $f_{2008}\left(f_{2008}(n)\right)$ as well. Actually, the latter is trivial if $f_{2008}(n)=n$; otherwise, $f_{2008}(n)=n+1$, which implies $2007 \nmid n+1$ and hence $n+1=f_{2008}(n+1)=f_{2008}\left(f_{2008}(n)\right)$.

So, if $2007 \mid m+n$, then

$$
f_{2008}(m+n)=m+n+1=(m+1)+(n+1)-1 \geq f_{2008}(m)+f_{2008}\left(f_{2008}(n)\right)-1 .
$$

Otherwise, $2007 \nmid m+n$, hence $2007 \nmid m$ or $2007 \nmid n$. In the former case we have $f_{2008}(m)=m$, while in the latter one $f_{2008}\left(f_{2008}(n)\right)=f_{2008}(n)=n$, providing

$$
f_{2008}(m)+f_{2008}\left(f_{2008}(n)\right)-1 \leq(m+n+1)-1=f_{2008}(m+n)
$$

Comment. The examples above are not unique. The values $1,2, \ldots, 2008$ can be realized in several ways. Here we present other two constructions for $j \leq 2007$, without proof:

$$
g_{j}(n)=\left\{\begin{array}{ll}
1, & n<2007, \\
j, & n=2007, \\
n, & n>2007 ;
\end{array} \quad h_{j}(n)=\max \left\{1,\left\lfloor\frac{j n}{2007}\right\rfloor\right\} .\right.
$$

Also the example for $j=2008$ can be generalized. In particular, choosing a divisor $d>1$ of 2007, one can set

$$
f_{2008, d}(n)= \begin{cases}n, & d \nmid n, \\ n+1, & d \mid n .\end{cases}
$$

A3. Let $n$ be a positive integer, and let $x$ and $y$ be positive real numbers such that $x^{n}+y^{n}=1$. Prove that

$$
\left(\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}\right)\left(\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}\right)<\frac{1}{(1-x)(1-y)} .
$$

(Estonia)
Solution 1. For each real $t \in(0,1)$,

$$
\frac{1+t^{2}}{1+t^{4}}=\frac{1}{t}-\frac{(1-t)\left(1-t^{3}\right)}{t\left(1+t^{4}\right)}<\frac{1}{t}
$$

Substituting $t=x^{k}$ and $t=y^{k}$,

$$
0<\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}<\sum_{k=1}^{n} \frac{1}{x^{k}}=\frac{1-x^{n}}{x^{n}(1-x)} \quad \text { and } \quad 0<\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}<\sum_{k=1}^{n} \frac{1}{y^{k}}=\frac{1-y^{n}}{y^{n}(1-y)}
$$

Since $1-y^{n}=x^{n}$ and $1-x^{n}=y^{n}$,

$$
\frac{1-x^{n}}{x^{n}(1-x)}=\frac{y^{n}}{x^{n}(1-x)}, \quad \frac{1-y^{n}}{y^{n}(1-y)}=\frac{x^{n}}{y^{n}(1-y)}
$$

and therefore

$$
\left(\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}\right)\left(\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}\right)<\frac{y^{n}}{x^{n}(1-x)} \cdot \frac{x^{n}}{y^{n}(1-y)}=\frac{1}{(1-x)(1-y)}
$$

Solution 2. We prove

$$
\begin{equation*}
\left(\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}\right)\left(\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}\right)<\frac{\left(\frac{1+\sqrt{2}}{2} \ln 2\right)^{2}}{(1-x)(1-y)}<\frac{0.7001}{(1-x)(1-y)} \tag{1}
\end{equation*}
$$

The idea is to estimate each term on the left-hand side with the same constant. To find the upper bound for the expression $\frac{1+x^{2 k}}{1+x^{4 k}}$, consider the function $f(t)=\frac{1+t}{1+t^{2}}$ in interval $(0,1)$. Since

$$
f^{\prime}(t)=\frac{1-2 t-t^{2}}{\left(1+t^{2}\right)^{2}}=\frac{(\sqrt{2}+1+t)(\sqrt{2}-1-t)}{\left(1+t^{2}\right)^{2}}
$$

the function increases in interval $(0, \sqrt{2}-1]$ and decreases in $[\sqrt{2}-1,1)$. Therefore the maximum is at point $t_{0}=\sqrt{2}-1$ and

$$
f(t)=\frac{1+t}{1+t^{2}} \leq f\left(t_{0}\right)=\frac{1+\sqrt{2}}{2}=\alpha .
$$

Applying this to each term on the left-hand side of (1), we obtain

$$
\begin{equation*}
\left(\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}\right)\left(\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}\right) \leq n \alpha \cdot n \alpha=(n \alpha)^{2} . \tag{2}
\end{equation*}
$$

To estimate $(1-x)(1-y)$ on the right-hand side, consider the function

$$
g(t)=\ln \left(1-t^{1 / n}\right)+\ln \left(1-(1-t)^{1 / n}\right) .
$$

Substituting $s$ for $1-t$, we have

$$
-n g^{\prime}(t)=\frac{t^{1 / n-1}}{1-t^{1 / n}}-\frac{s^{1 / n-1}}{1-s^{1 / n}}=\frac{1}{s t}\left(\frac{(1-t) t^{1 / n}}{1-t^{1 / n}}-\frac{(1-s) s^{1 / n}}{1-s^{1 / n}}\right)=\frac{h(t)-h(s)}{s t}
$$

The function

$$
h(t)=t^{1 / n} \frac{1-t}{1-t^{1 / n}}=\sum_{i=1}^{n} t^{i / n}
$$

is obviously increasing for $t \in(0,1)$, hence for these values of $t$ we have

$$
g^{\prime}(t)>0 \Longleftrightarrow h(t)<h(s) \Longleftrightarrow t<s=1-t \Longleftrightarrow t<\frac{1}{2} .
$$

Then, the maximum of $g(t)$ in $(0,1)$ is attained at point $t_{1}=1 / 2$ and therefore

$$
g(t) \leq g\left(\frac{1}{2}\right)=2 \ln \left(1-2^{-1 / n}\right), \quad t \in(0,1)
$$

Substituting $t=x^{n}$, we have $1-t=y^{n},(1-x)(1-y)=\exp g(t)$ and hence

$$
\begin{equation*}
(1-x)(1-y)=\exp g(t) \leq\left(1-2^{-1 / n}\right)^{2} \tag{3}
\end{equation*}
$$

Combining (2) and (3), we get

$$
\left(\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}\right)\left(\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}\right) \leq(\alpha n)^{2} \cdot 1 \leq(\alpha n)^{2} \frac{\left(1-2^{-1 / n}\right)^{2}}{(1-x)(1-y)}=\frac{\left(\alpha n\left(1-2^{-1 / n}\right)\right)^{2}}{(1-x)(1-y)} .
$$

Applying the inequality $1-\exp (-t)<t$ for $t=\frac{\ln 2}{n}$, we obtain

$$
\alpha n\left(1-2^{-1 / n}\right)=\alpha n\left(1-\exp \left(-\frac{\ln 2}{n}\right)\right)<\alpha n \cdot \frac{\ln 2}{n}=\alpha \ln 2=\frac{1+\sqrt{2}}{2} \ln 2 .
$$

Hence,

$$
\left(\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}\right)\left(\sum_{k=1}^{n} \frac{1+y^{2 k}}{1+y^{4 k}}\right)<\frac{\left(\frac{1+\sqrt{2}}{2} \ln 2\right)^{2}}{(1-x)(1-y)}
$$

Comment. It is a natural idea to compare the sum $S_{n}(x)=\sum_{k=1}^{n} \frac{1+x^{2 k}}{1+x^{4 k}}$ with the integral $I_{n}(x)=$ $\int_{0}^{n} \frac{1+x^{2 t}}{1+x^{4 t}} \mathrm{~d} t$. Though computing the integral is quite standard, many difficulties arise. First, the integrand $\frac{1+x^{2 k}}{1+x^{4 k}}$ has an increasing segment and, depending on $x$, it can have a decreasing segment as well. So comparing $S_{n}(x)$ and $I_{n}(x)$ is not completely obvious. We can add a term to fix the estimate, e.g. $S_{n} \leq I_{n}+(\alpha-1)$, but then the final result will be weak for the small values of $n$. Second, we have to minimize $(1-x)(1-y) I_{n}(x) I_{n}(y)$ which leads to very unpleasant computations.

However, by computer search we found that the maximum of $I_{n}(x) I_{n}(y)$ is at $x=y=2^{-1 / n}$, as well as the maximum of $S_{n}(x) S_{n}(y)$, and the latter is less. Hence, one can conjecture that the exact constant which can be put into the numerator on the right-hand side of (1) is

$$
\left(\ln 2 \cdot \int_{0}^{1} \frac{1+4^{-t}}{1+16^{-t}} \mathrm{~d} t\right)^{2}=\frac{1}{4}\left(\frac{1}{2} \ln \frac{17}{2}+\arctan 4-\frac{\pi}{4}\right)^{2} \approx 0.6484 .
$$

A4. Find all functions $f: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$such that

$$
\begin{equation*}
f(x+f(y))=f(x+y)+f(y) \tag{1}
\end{equation*}
$$

for all $x, y \in \mathbb{R}^{+}$. (Symbol $\mathbb{R}^{+}$denotes the set of all positive real numbers.)
(Thaliand)
Answer. $f(x)=2 x$.
Solution 1. First we show that $f(y)>y$ for all $y \in \mathbb{R}^{+}$. Functional equation (1) yields $f(x+f(y))>f(x+y)$ and hence $f(y) \neq y$ immediately. If $f(y)<y$ for some $y$, then setting $x=y-f(y)$ we get

$$
f(y)=f((y-f(y))+f(y))=f((y-f(y))+y)+f(y)>f(y)
$$

contradiction. Therefore $f(y)>y$ for all $y \in \mathbb{R}^{+}$.
For $x \in \mathbb{R}^{+}$define $g(x)=f(x)-x$; then $f(x)=g(x)+x$ and, as we have seen, $g(x)>0$. Transforming (1) for function $g(x)$ and setting $t=x+y$,

$$
\begin{aligned}
f(t+g(y)) & =f(t)+f(y) \\
g(t+g(y))+t+g(y) & =(g(t)+t)+(g(y)+y)
\end{aligned}
$$

and therefore

$$
\begin{equation*}
g(t+g(y))=g(t)+y \quad \text { for all } t>y>0 \tag{2}
\end{equation*}
$$

Next we prove that function $g(x)$ is injective. Suppose that $g\left(y_{1}\right)=g\left(y_{2}\right)$ for some numbers $y_{1}, y_{2} \in \mathbb{R}^{+}$. Then by (2),

$$
g(t)+y_{1}=g\left(t+g\left(y_{1}\right)\right)=g\left(t+g\left(y_{2}\right)\right)=g(t)+y_{2}
$$

for all $t>\max \left\{y_{1}, y_{2}\right\}$. Hence, $g\left(y_{1}\right)=g\left(y_{2}\right)$ is possible only if $y_{1}=y_{2}$.
Now let $u, v$ be arbitrary positive numbers and $t>u+v$. Applying (2) three times,

$$
g(t+g(u)+g(v))=g(t+g(u))+v=g(t)+u+v=g(t+g(u+v)) .
$$

By the injective property we conclude that $t+g(u)+g(v)=t+g(u+v)$, hence

$$
\begin{equation*}
g(u)+g(v)=g(u+v) . \tag{3}
\end{equation*}
$$

Since function $g(v)$ is positive, equation (3) also shows that $g$ is an increasing function.
Finally we prove that $g(x)=x$. Combining (2) and (3), we obtain

$$
g(t)+y=g(t+g(y))=g(t)+g(g(y))
$$

and hence

$$
g(g(y))=y
$$

Suppose that there exists an $x \in \mathbb{R}^{+}$such that $g(x) \neq x$. By the monotonicity of $g$, if $x>g(x)$ then $g(x)>g(g(x))=x$. Similarly, if $x<g(x)$ then $g(x)<g(g(x))=x$. Both cases lead to contradiction, so there exists no such $x$.

We have proved that $g(x)=x$ and therefore $f(x)=g(x)+x=2 x$ for all $x \in \mathbb{R}^{+}$. This function indeed satisfies the functional equation (1).

Comment. It is well-known that the additive property (3) together with $g(x) \geq 0$ (for $x>0$ ) imply $g(x)=c x$. So, after proving (3), it is sufficient to test functions $f(x)=(c+1) x$.
Solution 2. We prove that $f(y)>y$ and introduce function $g(x)=f(x)-x>0$ in the same way as in Solution 1.

For arbitrary $t>y>0$, substitute $x=t-y$ into (1) to obtain

$$
f(t+g(y))=f(t)+f(y)
$$

which, by induction, implies

$$
\begin{equation*}
f(t+n g(y))=f(t)+n f(y) \quad \text { for all } t>y>0, n \in \mathbb{N} . \tag{4}
\end{equation*}
$$

Take two arbitrary positive reals $y$ and $z$ and a third fixed number $t>\max \{y, z\}$. For each positive integer $k$, let $\ell_{k}=\left\lfloor k \frac{g(y)}{g(z)}\right\rfloor$. Then $t+k g(y)-\ell_{k} g(z) \geq t>z$ and, applying (4) twice,

$$
\begin{aligned}
f\left(t+k g(y)-\ell_{k} g(z)\right)+\ell_{k} f(z) & =f(t+k g(y))=f(t)+k f(y) \\
0<\frac{1}{k} f\left(t+k g(y)-\ell_{k} g(z)\right) & =\frac{f(t)}{k}+f(y)-\frac{\ell_{k}}{k} f(z)
\end{aligned}
$$

As $k \rightarrow \infty$ we get

$$
0 \leq \lim _{k \rightarrow \infty}\left(\frac{f(t)}{k}+f(y)-\frac{\ell_{k}}{k} f(z)\right)=f(y)-\frac{g(y)}{g(z)} f(z)=f(y)-\frac{f(y)-y}{f(z)-z} f(z)
$$

and therefore

$$
\frac{f(y)}{y} \leq \frac{f(z)}{z}
$$

Exchanging variables $y$ and $z$, we obtain the reverse inequality. Hence, $\frac{f(y)}{y}=\frac{f(z)}{z}$ for arbitrary $y$ and $z$; so function $\frac{f(x)}{x}$ is constant, $f(x)=c x$.

Substituting back into (1), we find that $f(x)=c x$ is a solution if and only if $c=2$. So the only solution for the problem is $f(x)=2 x$.

A5. Let $c>2$, and let $a(1), a(2), \ldots$ be a sequence of nonnegative real numbers such that

$$
\begin{equation*}
a(m+n) \leq 2 a(m)+2 a(n) \quad \text { for all } m, n \geq 1, \tag{1}
\end{equation*}
$$

and

$$
\begin{equation*}
a\left(2^{k}\right) \leq \frac{1}{(k+1)^{c}} \quad \text { for all } k \geq 0 \tag{2}
\end{equation*}
$$

Prove that the sequence $a(n)$ is bounded.
(Croatia)
Solution 1. For convenience, define $a(0)=0$; then condition (1) persists for all pairs of nonnegative indices.
Lemma 1. For arbitrary nonnegative indices $n_{1}, \ldots, n_{k}$, we have

$$
\begin{equation*}
a\left(\sum_{i=1}^{k} n_{i}\right) \leq \sum_{i=1}^{k} 2^{i} a\left(n_{i}\right) \tag{3}
\end{equation*}
$$

and

$$
\begin{equation*}
a\left(\sum_{i=1}^{k} n_{i}\right) \leq 2 k \sum_{i=1}^{k} a\left(n_{i}\right) . \tag{4}
\end{equation*}
$$

Proof. Inequality (3) is proved by induction on $k$. The base case $k=1$ is trivial, while the induction step is provided by
$a\left(\sum_{i=1}^{k+1} n_{i}\right)=a\left(n_{1}+\sum_{i=2}^{k+1} n_{i}\right) \leq 2 a\left(n_{1}\right)+2 a\left(\sum_{i=1}^{k} n_{i+1}\right) \leq 2 a\left(n_{1}\right)+2 \sum_{i=1}^{k} 2^{i} a\left(n_{i+1}\right)=\sum_{i=1}^{k+1} 2^{i} a\left(n_{i}\right)$.
To establish (4), first the inequality

$$
a\left(\sum_{i=1}^{2^{d}} n_{i}\right) \leq 2^{d} \sum_{i=1}^{2^{d}} a\left(n_{i}\right)
$$

can be proved by an obvious induction on $d$. Then, turning to (4), we find an integer $d$ such that $2^{d-1}<k \leq 2^{d}$ to obtain

$$
a\left(\sum_{i=1}^{k} n_{i}\right)=a\left(\sum_{i=1}^{k} n_{i}+\sum_{i=k+1}^{2^{d}} 0\right) \leq 2^{d}\left(\sum_{i=1}^{k} a\left(n_{i}\right)+\sum_{i=k+1}^{2^{d}} a(0)\right)=2^{d} \sum_{i=1}^{k} a\left(n_{i}\right) \leq 2 k \sum_{i=1}^{k} a\left(n_{i}\right) .
$$

Fix an increasing unbounded sequence $0=M_{0}<M_{1}<M_{2}<\ldots$ of real numbers; the exact values will be defined later. Let $n$ be an arbitrary positive integer and write

$$
n=\sum_{i=0}^{d} \varepsilon_{i} \cdot 2^{i}, \quad \text { where } \varepsilon_{i} \in\{0,1\}
$$

Set $\varepsilon_{i}=0$ for $i>d$, and take some positive integer $f$ such that $M_{f}>d$. Applying (3), we get

$$
a(n)=a\left(\sum_{k=1}^{f} \sum_{M_{k-1} \leq i<M_{k}} \varepsilon_{i} \cdot 2^{i}\right) \leq \sum_{k=1}^{f} 2^{k} a\left(\sum_{M_{k-1} \leq i<M_{k}} \varepsilon_{i} \cdot 2^{i}\right) .
$$

Note that there are less than $M_{k}-M_{k-1}+1$ integers in interval $\left[M_{k-1}, M_{k}\right.$ ); hence, using (4) we have

$$
\begin{aligned}
a(n) & \leq \sum_{k=1}^{f} 2^{k} \cdot 2\left(M_{k}-M_{k-1}+1\right) \sum_{M_{k-1} \leq i<M_{k}} \varepsilon_{i} \cdot a\left(2^{i}\right) \\
& \leq \sum_{k=1}^{f} 2^{k} \cdot 2\left(M_{k}-M_{k-1}+1\right)^{2} \max _{M_{k-1} \leq i<M_{k}} a\left(2^{i}\right) \\
& \leq \sum_{k=1}^{f} 2^{k+1}\left(M_{k}+1\right)^{2} \cdot \frac{1}{\left(M_{k-1}+1\right)^{c}}=\sum_{k=1}^{f}\left(\frac{M_{k}+1}{M_{k-1}+1}\right)^{2} \frac{2^{k+1}}{\left(M_{k-1}+1\right)^{c-2}} .
\end{aligned}
$$

Setting $M_{k}=4^{k /(c-2)}-1$, we obtain

$$
a(n) \leq \sum_{k=1}^{f} 4^{2 /(c-2)} \frac{2^{k+1}}{\left(4^{(k-1) /(c-2)}\right)^{c-2}}=8 \cdot 4^{2 /(c-2)} \sum_{k=1}^{f}\left(\frac{1}{2}\right)^{k}<8 \cdot 4^{2 /(c-2)},
$$

and the sequence $a(n)$ is bounded.

## Solution 2.

Lemma 2. Suppose that $s_{1}, \ldots, s_{k}$ are positive integers such that

$$
\sum_{i=1}^{k} 2^{-s_{i}} \leq 1
$$

Then for arbitrary positive integers $n_{1}, \ldots, n_{k}$ we have

$$
a\left(\sum_{i=1}^{k} n_{i}\right) \leq \sum_{i=1}^{k} 2^{s_{i}} a\left(n_{i}\right)
$$

Proof. Apply an induction on $k$. The base cases are $k=1$ (trivial) and $k=2$ (follows from the condition (1)). Suppose that $k>2$. We can assume that $s_{1} \leq s_{2} \leq \cdots \leq s_{k}$. Note that

$$
\sum_{i=1}^{k-1} 2^{-s_{i}} \leq 1-2^{-s_{k-1}}
$$

since the left-hand side is a fraction with the denominator $2^{s_{k-1}}$, and this fraction is less than 1. Define $s_{k-1}^{\prime}=s_{k-1}-1$ and $n_{k-1}^{\prime}=n_{k-1}+n_{k}$; then we have

$$
\sum_{i=1}^{k-2} 2^{-s_{i}}+2^{-s_{k-1}^{\prime}} \leq\left(1-2 \cdot 2^{-s_{k-1}}\right)+2^{1-s_{k-1}}=1
$$

Now, the induction hypothesis can be applied to achieve

$$
\begin{aligned}
a\left(\sum_{i=1}^{k} n_{i}\right)=a\left(\sum_{i=1}^{k-2} n_{i}+n_{k-1}^{\prime}\right) & \leq \sum_{i=1}^{k-2} 2^{s_{i}} a\left(n_{i}\right)+2^{s_{k-1}^{\prime}} a\left(n_{k-1}^{\prime}\right) \\
& \leq \sum_{i=1}^{k-2} 2^{s_{i}} a\left(n_{i}\right)+2^{s_{k-1}-1} \cdot 2\left(a\left(n_{k-1}\right)+a\left(n_{k}\right)\right) \\
& \leq \sum_{i=1}^{k-2} 2^{s_{i}} a\left(n_{i}\right)+2^{s_{k-1}} a\left(n_{k-1}\right)+2^{s_{k}} a\left(n_{k}\right)
\end{aligned}
$$

Let $q=c / 2>1$. Take an arbitrary positive integer $n$ and write

$$
n=\sum_{i=1}^{k} 2^{u_{i}}, \quad 0 \leq u_{1}<u_{2}<\cdots<u_{k}
$$

Choose $s_{i}=\left\lfloor\log _{2}\left(u_{i}+1\right)^{q}\right\rfloor+d(i=1, \ldots, k)$ for some integer $d$. We have

$$
\sum_{i=1}^{k} 2^{-s_{i}}=2^{-d} \sum_{i=1}^{k} 2^{-\left\lfloor\log _{2}\left(u_{i}+1\right)^{q}\right\rfloor}
$$

and we choose $d$ in such a way that

$$
\frac{1}{2}<\sum_{i=1}^{k} 2^{-s_{i}} \leq 1
$$

In particular, this implies

$$
2^{d}<2 \sum_{i=1}^{k} 2^{-\left\lfloor\log _{2}\left(u_{i}+1\right)^{q}\right\rfloor}<4 \sum_{i=1}^{k} \frac{1}{\left(u_{i}+1\right)^{q}} .
$$

Now, by Lemma 2 we obtain

$$
\begin{aligned}
a(n)=a\left(\sum_{i=1}^{k} 2^{u_{i}}\right) & \leq \sum_{i=1}^{k} 2^{s_{i}} a\left(2^{u_{i}}\right) \leq \sum_{i=1}^{k} 2^{d}\left(u_{i}+1\right)^{q} \cdot \frac{1}{\left(u_{i}+1\right)^{2 q}} \\
& =2^{d} \sum_{i=1}^{k} \frac{1}{\left(u_{i}+1\right)^{q}}<4\left(\sum_{i=1}^{k} \frac{1}{\left(u_{i}+1\right)^{q}}\right)^{2},
\end{aligned}
$$

which is bounded since $q>1$.
Comment 1. In fact, Lemma 2 (applied to the case $n_{i}=2^{u_{i}}$ only) provides a sharp bound for any $a(n)$. Actually, let $b(k)=\frac{1}{(k+1)^{c}}$ and consider the sequence

$$
\begin{equation*}
a(n)=\min \left\{\sum_{i=1}^{k} 2^{s_{i}} b\left(u_{i}\right) \mid k \in \mathbb{N}, \quad \sum_{i=1}^{k} 2^{-s_{i}} \leq 1, \quad \sum_{i=1}^{k} 2^{u_{i}}=n\right\} . \tag{5}
\end{equation*}
$$

We show that this sequence satisfies the conditions of the problem. Take two arbitrary indices $m$ and $n$. Let

$$
\begin{aligned}
& a(m)=\sum_{i=1}^{k} 2^{s_{i}} b\left(u_{i}\right), \quad \sum_{i=1}^{k} 2^{-s_{i}} \leq 1, \quad \sum_{i=1}^{k} 2^{u_{i}}=m ; \\
& a(n)=\sum_{i=1}^{l} 2^{r_{i}} b\left(w_{i}\right), \quad \sum_{i=1}^{l} 2^{-r_{i}} \leq 1, \quad \sum_{i=1}^{l} 2^{w_{i}}=n .
\end{aligned}
$$

Then we have

$$
\sum_{i=1}^{k} 2^{-1-s_{i}}+\sum_{i=1}^{l} 2^{-1-r_{i}} \leq \frac{1}{2}+\frac{1}{2}=1, \quad \sum_{i=1}^{k} 2^{u_{i}}+\sum_{i=1}^{l} 2^{w_{i}}=m+n,
$$

so by (5) we obtain

$$
a(n+m) \leq \sum_{i=1}^{k} 2^{1+s_{i}} b\left(u_{i}\right)+\sum_{i=1}^{l} 2^{1+r_{i}} b\left(w_{i}\right)=2 a(m)+2 a(n) .
$$

Comment 2. The condition $c>2$ is sharp; we show that the sequence (5) is not bounded if $c \leq 2$.
First, we prove that for an arbitrary $n$ the minimum in (5) is attained with a sequence ( $u_{i}$ ) consisting of distinct numbers. To the contrary, assume that $u_{k-1}=u_{k}$. Replace $u_{k-1}$ and $u_{k}$ by a single number $u_{k-1}^{\prime}=u_{k}+1$, and $s_{k-1}$ and $s_{k}$ by $s_{k-1}^{\prime}=\min \left\{s_{k-1}, s_{k}\right\}$. The modified sequences provide a better bound since

$$
2^{s_{k-1}^{\prime}} b\left(u_{k-1}^{\prime}\right)=2^{s_{k-1}^{\prime}} b\left(u_{k}+1\right)<2^{s_{k-1}} b\left(u_{k-1}\right)+2^{s_{k}} b\left(u_{k}\right)
$$

(we used the fact that $b(k)$ is decreasing). This is impossible.
Hence, the claim is proved, and we can assume that the minimum is attained with $u_{1}<\cdots<u_{k}$; then

$$
n=\sum_{i=1}^{k} 2^{u_{i}}
$$

is simply the binary representation of $n$. (In particular, it follows that $a\left(2^{n}\right)=b(n)$ for each $n$.)
Now we show that the sequence $\left(a\left(2^{k}-1\right)\right)$ is not bounded. For some $s_{1}, \ldots, s_{k}$ we have

$$
a\left(2^{k}-1\right)=a\left(\sum_{i=1}^{k} 2^{i-1}\right)=\sum_{i=1}^{k} 2^{s_{i}} b(i-1)=\sum_{i=1}^{k} \frac{2^{s_{i}}}{i^{c}} .
$$

By the Cauchy-Schwarz inequality we get

$$
a\left(2^{k}-1\right)=a\left(2^{k}-1\right) \cdot 1 \geq\left(\sum_{i=1}^{k} \frac{2^{s_{i}}}{i^{c}}\right)\left(\sum_{i=1}^{k} \frac{1}{2^{s_{i}}}\right) \geq\left(\sum_{i=1}^{k} \frac{1}{i^{c / 2}}\right)^{2},
$$

which is unbounded.
For $c \leq 2$, it is also possible to show a concrete counterexample. Actually, one can prove that the sequence

$$
a\left(\sum_{i=1}^{k} 2^{u_{i}}\right)=\sum_{i=1}^{k} \frac{i}{\left(u_{i}+1\right)^{2}} \quad\left(0 \leq u_{1}<\ldots<u_{k}\right)
$$

satisfies (1) and (2) but is not bounded.

A6. Let $a_{1}, a_{2}, \ldots, a_{100}$ be nonnegative real numbers such that $a_{1}^{2}+a_{2}^{2}+\ldots+a_{100}^{2}=1$. Prove that

$$
a_{1}^{2} a_{2}+a_{2}^{2} a_{3}+\ldots+a_{100}^{2} a_{1}<\frac{12}{25}
$$

(Poland)
Solution. Let $S=\sum_{k=1}^{100} a_{k}^{2} a_{k+1}$. (As usual, we consider the indices modulo 100, e.g. we set $a_{101}=a_{1}$ and $a_{102}=a_{2}$.)

Applying the Cauchy-Schwarz inequality to sequences $\left(a_{k+1}\right)$ and $\left(a_{k}^{2}+2 a_{k+1} a_{k+2}\right)$, and then the AM-GM inequality to numbers $a_{k+1}^{2}$ and $a_{k+2}^{2}$,

$$
\begin{align*}
(3 S)^{2} & =\left(\sum_{k=1}^{100} a_{k+1}\left(a_{k}^{2}+2 a_{k+1} a_{k+2}\right)\right)^{2} \leq\left(\sum_{k=1}^{100} a_{k+1}^{2}\right)\left(\sum_{k=1}^{100}\left(a_{k}^{2}+2 a_{k+1} a_{k+2}\right)^{2}\right)  \tag{1}\\
& =1 \cdot \sum_{k=1}^{100}\left(a_{k}^{2}+2 a_{k+1} a_{k+2}\right)^{2}=\sum_{k=1}^{100}\left(a_{k}^{4}+4 a_{k}^{2} a_{k+1} a_{k+2}+4 a_{k+1}^{2} a_{k+2}^{2}\right) \\
& \leq \sum_{k=1}^{100}\left(a_{k}^{4}+2 a_{k}^{2}\left(a_{k+1}^{2}+a_{k+2}^{2}\right)+4 a_{k+1}^{2} a_{k+2}^{2}\right)=\sum_{k=1}^{100}\left(a_{k}^{4}+6 a_{k}^{2} a_{k+1}^{2}+2 a_{k}^{2} a_{k+2}^{2}\right) .
\end{align*}
$$

Applying the trivial estimates

$$
\sum_{k=1}^{100}\left(a_{k}^{4}+2 a_{k}^{2} a_{k+1}^{2}+2 a_{k}^{2} a_{k+2}^{2}\right) \leq\left(\sum_{k=1}^{100} a_{k}^{2}\right)^{2} \quad \text { and } \quad \sum_{k=1}^{100} a_{k}^{2} a_{k+1}^{2} \leq\left(\sum_{i=1}^{50} a_{2 i-1}^{2}\right)\left(\sum_{j=1}^{50} a_{2 j}^{2}\right)
$$

we obtain that

$$
(3 S)^{2} \leq\left(\sum_{k=1}^{100} a_{k}^{2}\right)^{2}+4\left(\sum_{i=1}^{50} a_{2 i-1}^{2}\right)\left(\sum_{j=1}^{50} a_{2 j}^{2}\right) \leq 1+\left(\sum_{i=1}^{50} a_{2 i-1}^{2}+\sum_{j=1}^{50} a_{2 j}^{2}\right)^{2}=2
$$

hence

$$
S \leq \frac{\sqrt{2}}{3} \approx 0.4714<\frac{12}{25}=0.48
$$

Comment 1. By applying the Lagrange multiplier method, one can see that the maximum is attained at values of $a_{i}$ satisfying

$$
\begin{equation*}
a_{k-1}^{2}+2 a_{k} a_{k+1}=2 \lambda a_{k} \tag{2}
\end{equation*}
$$

for all $k=1,2, \ldots, 100$. Though this system of equations seems hard to solve, it can help to find the estimate above; it may suggest to have a closer look at the expression $a_{k-1}^{2} a_{k}+2 a_{k}^{2} a_{k+1}$.

Moreover, if the numbers $a_{1}, \ldots, a_{100}$ satisfy (2), we have equality in (1). (See also Comment 3.)
Comment 2. It is natural to ask what is the best constant $c_{n}$ in the inequality

$$
\begin{equation*}
a_{1}^{2} a_{2}+a_{2}^{2} a_{3}+\ldots+a_{n}^{2} a_{1} \leq c_{n}\left(a_{1}^{2}+a_{2}^{2}+\ldots+a_{n}^{2}\right)^{3 / 2} \tag{3}
\end{equation*}
$$

For $1 \leq n \leq 4$ one may prove $c_{n}=1 / \sqrt{n}$ which is achieved when $a_{1}=a_{2}=\ldots=a_{n}$. However, the situation changes completely if $n \geq 5$. In this case we do not know the exact value of $c_{n}$. By computer search it can be found that $c_{n} \approx 0.4514$ and it is realized for example if

$$
a_{1} \approx 0.5873, \quad a_{2} \approx 0.6771, \quad a_{3} \approx 0.4224, \quad a_{4} \approx 0.1344, \quad a_{5} \approx 0.0133
$$

and $a_{k} \approx 0$ for $k \geq 6$. This example also proves that $c_{n}>0.4513$.

Comment 3. The solution can be improved in several ways to give somewhat better bounds for $c_{n}$. Here we show a variant which proves $c_{n}<0.4589$ for $n \geq 5$.

The value of $c_{n}$ does not change if negative values are also allowed in (3). So the problem is equivalent to maximizing

$$
f\left(a_{1}, a_{2}, \ldots, a_{n}\right)=a_{1}^{2} a_{2}+a_{2}^{2} a_{3}+\ldots+a_{n}^{2} a_{1}
$$

on the unit sphere $a_{1}^{2}+a_{2}^{2}+\ldots+a_{n}^{2}=1$ in $\mathbb{R}^{n}$. Since the unit sphere is compact, the function has a maximum and we can apply the Lagrange multiplier method; for each maximum point there exists a real number $\lambda$ such that

$$
a_{k-1}^{2}+2 a_{k} a_{k+1}=\lambda \cdot 2 a_{k} \quad \text { for all } k=1,2, \ldots, n .
$$

Then

$$
3 S=\sum_{k=1}^{n}\left(a_{k-1}^{2} a_{k}+2 a_{k}^{2} a_{k+1}\right)=\sum_{k=1}^{n} 2 \lambda a_{k}^{2}=2 \lambda
$$

and therefore

$$
\begin{equation*}
a_{k-1}^{2}+2 a_{k} a_{k+1}=3 S a_{k} \quad \text { for all } k=1,2, \ldots, n \tag{4}
\end{equation*}
$$

From (4) we can derive

$$
\begin{equation*}
9 S^{2}=\sum_{k=1}^{n}\left(3 S a_{k}\right)^{2}=\sum_{k=1}^{n}\left(a_{k-1}^{2}+2 a_{k} a_{k+1}\right)^{2}=\sum_{k=1}^{n} a_{k}^{4}+4 \sum_{k=1}^{n} a_{k}^{2} a_{k+1}^{2}+4 \sum_{k=1}^{n} a_{k}^{2} a_{k+1} a_{k+2} \tag{5}
\end{equation*}
$$

and

$$
\begin{equation*}
3 S^{2}=\sum_{k=1}^{n} 3 S a_{k-1}^{2} a_{k}=\sum_{k=1}^{n} a_{k-1}^{2}\left(a_{k-1}^{2}+2 a_{k} a_{k+1}\right)=\sum_{k=1}^{n} a_{k}^{4}+2 \sum_{k=1}^{n} a_{k}^{2} a_{k+1} a_{k+2} . \tag{6}
\end{equation*}
$$

Let $p$ be a positive number. Combining (5) and (6) and applying the AM-GM inequality,

$$
\begin{aligned}
(9+3 p) S^{2} & =(1+p) \sum_{k=1}^{n} a_{k}^{4}+4 \sum_{k=1}^{n} a_{k}^{2} a_{k+1}^{2}+(4+2 p) \sum_{k=1}^{n} a_{k}^{2} a_{k+1} a_{k+2} \\
& \leq(1+p) \sum_{k=1}^{n} a_{k}^{4}+4 \sum_{k=1}^{n} a_{k}^{2} a_{k+1}^{2}+\sum_{k=1}^{n}\left(2(1+p) a_{k}^{2} a_{k+2}^{2}+\frac{(2+p)^{2}}{2(1+p)} a_{k}^{2} a_{k+1}^{2}\right) \\
& =(1+p) \sum_{k=1}^{n}\left(a_{k}^{4}+2 a_{k}^{2} a_{k+1}^{2}+2 a_{k}^{2} a_{k+2}^{2}\right)+\left(4+\frac{(2+p)^{2}}{2(1+p)}-2(1+p)\right) \sum_{k=1}^{n} a_{k}^{2} a_{k+1}^{2} \\
& \leq(1+p)\left(\sum_{k=1}^{n} a_{k}^{2}\right)^{2}+\frac{8+4 p-3 p^{2}}{2(1+p)} \sum_{k=1}^{n} a_{k}^{2} a_{k+1}^{2} \\
& =(1+p)+\frac{8+4 p-3 p^{2}}{2(1+p)} \sum_{k=1}^{n} a_{k}^{2} a_{k+1}^{2} .
\end{aligned}
$$

Setting $p=\frac{2+2 \sqrt{7}}{3}$ which is the positive root of $8+4 p-3 p^{2}=0$, we obtain

$$
S \leq \sqrt{\frac{1+p}{9+3 p}}=\sqrt{\frac{5+2 \sqrt{7}}{33+6 \sqrt{7}}} \approx 0.458879
$$

A7. Let $n>1$ be an integer. In the space, consider the set

$$
S=\{(x, y, z) \mid x, y, z \in\{0,1, \ldots, n\}, x+y+z>0\} .
$$

Find the smallest number of planes that jointly contain all $(n+1)^{3}-1$ points of $S$ but none of them passes through the origin.
(Netherlands)
Answer. $3 n$ planes.
Solution. It is easy to find $3 n$ such planes. For example, planes $x=i, y=i$ or $z=i$ $(i=1,2, \ldots, n)$ cover the set $S$ but none of them contains the origin. Another such collection consists of all planes $x+y+z=k$ for $k=1,2, \ldots, 3 n$.

We show that $3 n$ is the smallest possible number.
Lemma 1. Consider a nonzero polynomial $P\left(x_{1}, \ldots, x_{k}\right)$ in $k$ variables. Suppose that $P$ vanishes at all points $\left(x_{1}, \ldots, x_{k}\right)$ such that $x_{1}, \ldots, x_{k} \in\{0,1, \ldots, n\}$ and $x_{1}+\cdots+x_{k}>0$, while $P(0,0, \ldots, 0) \neq 0$. Then $\operatorname{deg} P \geq k n$.
Proof. We use induction on $k$. The base case $k=0$ is clear since $P \neq 0$. Denote for clarity $y=x_{k}$.

Let $R\left(x_{1}, \ldots, x_{k-1}, y\right)$ be the residue of $P$ modulo $Q(y)=y(y-1) \ldots(y-n)$. Polynomial $Q(y)$ vanishes at each $y=0,1, \ldots, n$, hence $P\left(x_{1}, \ldots, x_{k-1}, y\right)=R\left(x_{1}, \ldots, x_{k-1}, y\right)$ for all $x_{1}, \ldots, x_{k-1}, y \in\{0,1, \ldots, n\}$. Therefore, $R$ also satisfies the condition of the Lemma; moreover, $\operatorname{deg}_{y} R \leq n$. Clearly, $\operatorname{deg} R \leq \operatorname{deg} P$, so it suffices to prove that $\operatorname{deg} R \geq n k$.

Now, expand polynomial $R$ in the powers of $y$ :

$$
R\left(x_{1}, \ldots, x_{k-1}, y\right)=R_{n}\left(x_{1}, \ldots, x_{k-1}\right) y^{n}+R_{n-1}\left(x_{1}, \ldots, x_{k-1}\right) y^{n-1}+\cdots+R_{0}\left(x_{1}, \ldots, x_{k-1}\right) .
$$

We show that polynomial $R_{n}\left(x_{1}, \ldots, x_{k-1}\right)$ satisfies the condition of the induction hypothesis.
Consider the polynomial $T(y)=R(0, \ldots, 0, y)$ of degree $\leq n$. This polynomial has $n$ roots $y=1, \ldots, n$; on the other hand, $T(y) \not \equiv 0$ since $T(0) \neq 0$. Hence $\operatorname{deg} T=n$, and its leading coefficient is $R_{n}(0,0, \ldots, 0) \neq 0$. In particular, in the case $k=1$ we obtain that coefficient $R_{n}$ is nonzero.

Similarly, take any numbers $a_{1}, \ldots, a_{k-1} \in\{0,1, \ldots, n\}$ with $a_{1}+\cdots+a_{k-1}>0$. Substituting $x_{i}=a_{i}$ into $R\left(x_{1}, \ldots, x_{k-1}, y\right)$, we get a polynomial in $y$ which vanishes at all points $y=0, \ldots, n$ and has degree $\leq n$. Therefore, this polynomial is null, hence $R_{i}\left(a_{1}, \ldots, a_{k-1}\right)=0$ for all $i=0,1, \ldots, n$. In particular, $R_{n}\left(a_{1}, \ldots, a_{k-1}\right)=0$.

Thus, the polynomial $R_{n}\left(x_{1}, \ldots, x_{k-1}\right)$ satisfies the condition of the induction hypothesis. So, we have $\operatorname{deg} R_{n} \geq(k-1) n$ and $\operatorname{deg} P \geq \operatorname{deg} R \geq \operatorname{deg} R_{n}+n \geq k n$.

Now we can finish the solution. Suppose that there are $N$ planes covering all the points of $S$ but not containing the origin. Let their equations be $a_{i} x+b_{i} y+c_{i} z+d_{i}=0$. Consider the polynomial

$$
P(x, y, z)=\prod_{i=1}^{N}\left(a_{i} x+b_{i} y+c_{i} z+d_{i}\right)
$$

It has total degree $N$. This polynomial has the property that $P\left(x_{0}, y_{0}, z_{0}\right)=0$ for any $\left(x_{0}, y_{0}, z_{0}\right) \in S$, while $P(0,0,0) \neq 0$. Hence by Lemma 1 we get $N=\operatorname{deg} P \geq 3 n$, as desired.

Comment 1. There are many other collections of $3 n$ planes covering the set $S$ but not covering the origin.

Solution 2. We present a different proof of the main Lemma 1. Here we confine ourselves to the case $k=3$, which is applied in the solution, and denote the variables by $x, y$ and $z$. (The same proof works for the general statement as well.)

The following fact is known with various proofs; we provide one possible proof for the completeness.
Lemma 2. For arbitrary integers $0 \leq m<n$ and for an arbitrary polynomial $P(x)$ of degree $m$,

$$
\begin{equation*}
\sum_{k=0}^{n}(-1)^{k}\binom{n}{k} P(k)=0 \tag{1}
\end{equation*}
$$

Proof. We use an induction on $n$. If $n=1$, then $P(x)$ is a constant polynomial, hence $P(1)-P(0)=0$, and the base is proved.

For the induction step, define $P_{1}(x)=P(x+1)-P(x)$. Then clearly $\operatorname{deg} P_{1}=\operatorname{deg} P-1=$ $m-1<n-1$, hence by the induction hypothesis we get

$$
\begin{aligned}
0 & =-\sum_{k=0}^{n-1}(-1)^{k}\binom{n-1}{k} P_{1}(k)=\sum_{k=0}^{n-1}(-1)^{k}\binom{n-1}{k}(P(k)-P(k+1)) \\
& =\sum_{k=0}^{n-1}(-1)^{k}\binom{n-1}{k} P(k)-\sum_{k=0}^{n-1}(-1)^{k}\binom{n-1}{k} P(k+1) \\
& =\sum_{k=0}^{n-1}(-1)^{k}\binom{n-1}{k} P(k)+\sum_{k=1}^{n}(-1)^{k}\binom{n-1}{k-1} P(k) \\
& =P(0)+\sum_{k=1}^{n-1}(-1)^{k}\left(\binom{n-1}{k-1}+\binom{n-1}{k}\right) P(k)+(-1)^{n} P(n)=\sum_{k=0}^{n}(-1)^{k}\binom{n}{k} P(k) .
\end{aligned}
$$

Now return to the proof of Lemma 1. Suppose, to the contrary, that $\operatorname{deg} P=N<3 n$. Consider the sum

$$
\Sigma=\sum_{i=0}^{n} \sum_{j=0}^{n} \sum_{k=0}^{n}(-1)^{i+j+k}\binom{n}{i}\binom{n}{j}\binom{n}{k} P(i, j, k)
$$

The only nonzero term in this sum is $P(0,0,0)$ and its coefficient is $\binom{n}{0}^{3}=1$; therefore $\Sigma=P(0,0,0) \neq 0$.

On the other hand, if $P(x, y, z)=\sum_{\alpha+\beta+\gamma \leq N} p_{\alpha, \beta, \gamma} x^{\alpha} y^{\beta} z^{\gamma}$, then

$$
\begin{aligned}
\Sigma & =\sum_{i=0}^{n} \sum_{j=0}^{n} \sum_{k=0}^{n}(-1)^{i+j+k}\binom{n}{i}\binom{n}{j}\binom{n}{k} \sum_{\alpha+\beta+\gamma \leq N} p_{\alpha, \beta, \gamma} i^{\alpha} j^{\beta} k^{\gamma} \\
& =\sum_{\alpha+\beta+\gamma \leq N} p_{\alpha, \beta, \gamma}\left(\sum_{i=0}^{n}(-1)^{i}\binom{n}{i} i^{\alpha}\right)\left(\sum_{j=0}^{n}(-1)^{j}\binom{n}{j} j^{\beta}\right)\left(\sum_{k=0}^{n}(-1)^{k}\binom{n}{k} k^{\gamma}\right) .
\end{aligned}
$$

Consider an arbitrary term in this sum. We claim that it is zero. Since $N<3 n$, one of three inequalities $\alpha<n, \beta<n$ or $\gamma<n$ is valid. For the convenience, suppose that $\alpha<n$. Applying Lemma 2 to polynomial $x^{\alpha}$, we get $\sum_{i=0}^{n}(-1)^{i}\binom{n}{i} i^{\alpha}=0$, hence the term is zero as required.

This yields $\Sigma=0$ which is a contradiction. Therefore, $\operatorname{deg} P \geq 3 n$.

Comment 2. The proof does not depend on the concrete coefficients in Lemma 2. Instead of this Lemma, one can simply use the fact that there exist numbers $\alpha_{0}, \alpha_{1}, \ldots, \alpha_{n}\left(\alpha_{0} \neq 0\right)$ such that

$$
\sum_{k=0}^{n} \alpha_{k} k^{m}=0 \quad \text { for every } 0 \leq m<n
$$

This is a system of homogeneous linear equations in variables $\alpha_{i}$. Since the number of equations is less than the number of variables, the only nontrivial thing is that there exists a solution with $\alpha_{0} \neq 0$. It can be shown in various ways.

## Combinatorics

C1. Let $n>1$ be an integer. Find all sequences $a_{1}, a_{2}, \ldots, a_{n^{2}+n}$ satisfying the following conditions:
(a) $a_{i} \in\{0,1\}$ for all $1 \leq i \leq n^{2}+n$;
(b) $a_{i+1}+a_{i+2}+\ldots+a_{i+n}<a_{i+n+1}+a_{i+n+2}+\ldots+a_{i+2 n}$ for all $0 \leq i \leq n^{2}-n$.
(Serbia)
Answer. Such a sequence is unique. It can be defined as follows:

$$
a_{u+v n}=\left\{\begin{array}{ll}
0, & u+v \leq n,  \tag{1}\\
1, & u+v \geq n+1
\end{array} \quad \text { for all } 1 \leq u \leq n \text { and } 0 \leq v \leq n\right.
$$

The terms can be arranged into blocks of length $n$ as

$$
(\underbrace{0 \ldots 0}_{n})(\underbrace{0 \ldots 0}_{n-1} 1)(\underbrace{0 \ldots 0}_{n-2} 11) \ldots(\underbrace{0 \ldots 0}_{n-v} \underbrace{1 \ldots 1}_{v}) \ldots(0 \underbrace{1 \ldots 1}_{n-1})(\underbrace{1 \ldots 1}_{n}) .
$$

Solution 1. Consider a sequence $\left(a_{i}\right)$ satisfying the conditions. For arbitrary integers $0 \leq$ $k \leq l \leq n^{2}+n$ denote $S(k, l]=a_{k+1}+\cdots+a_{l}$. (If $k=l$ then $S(k, l]=0$.) Then condition (b) can be rewritten as $S(i, i+n]<S(i+n, i+2 n]$ for all $0 \leq i \leq n^{2}-n$. Notice that for $0 \leq k \leq l \leq m \leq n^{2}+n$ we have $S(k, m]=S(k, l]+S(l, m]$.

By condition (b),

$$
0 \leq S(0, n]<S(n, 2 n]<\cdots<S\left(n^{2}, n^{2}+n\right] \leq n
$$

We have only $n+1$ distinct integers in the interval $[0, n]$; hence,

$$
\begin{equation*}
S(v n,(v+1) n]=v \quad \text { for all } 0 \leq v \leq n . \tag{2}
\end{equation*}
$$

In particular, $S(0, n]=0$ and $S\left(n^{2}, n^{2}+n\right]=n$, therefore

$$
\begin{align*}
a_{1} & =a_{2}=\ldots=a_{n}=0,  \tag{3}\\
a_{n^{2}+1} & =a_{n^{2}+2}=\ldots=a_{n^{2}+n}=1 . \tag{4}
\end{align*}
$$

Subdivide sequence $\left(a_{i}\right)$ into $n+1$ blocks, each consisting of $n$ consecutive terms, and number them from 0 to $n$. We show by induction on $v$ that the $v$ th blocks has the form

$$
(\underbrace{0 \ldots 0}_{n-v} \underbrace{1 \ldots 1}_{v}) .
$$

The base case $v=0$ is provided by (3).

Consider the $v$ th block for $v>0$. By (2), it contains some "ones". Let the first "one" in this block be at the $u$ th position (that is, $a_{u+v n}=1$ ). By the induction hypothesis, the $(v-1)$ th and $v$ th blocks of $\left(a_{i}\right)$ have the form

$$
(\underbrace{0 \ldots \underbrace{0 \ldots 0}_{v-1} \underbrace{1 \ldots 1}_{u-1})(\underbrace{0 \ldots 0}_{u \ldots 0} 1}_{n-v+1} * \ldots *),
$$

where each star can appear to be any binary digit. Observe that $u \leq n-v+1$, since the sum in this block is $v$. Then, the fragment of length $n$ bracketed above has exactly $(v-1)+1$ ones, i. e. $S(u+(v-1) n, u+v n]=v$. Hence,

$$
v=S(u+(v-1) n, u+v n]<S(u+v n, u+(v+1) n]<\cdots<S\left(u+(n-1) n, u+n^{2}\right] \leq n
$$

we have $n-v+1$ distinct integers in the interval $[v, n]$, therefore $S(u+(t-1) n, u+t n]=t$ for each $t=v, \ldots, n$.

Thus, the end of sequence $\left(a_{i}\right)$ looks as following:

$$
(\underbrace{u \text { zeroes }}_{\sum=v-1} \overbrace{0 \ldots 01 \ldots 1}^{\sum=v})(\underbrace{0 \ldots 01 * \ldots *}_{\sum=v})(\underbrace{\sum \ldots * * \ldots *}_{\sum=v+1}) \cdots \overbrace{\underbrace{* \ldots 1}_{\sum}}^{\sum=v+1} \overbrace{\underbrace{1 \ldots 1}_{\sum=n}}^{\sum=n})
$$

(each bracketed fragment contains $n$ terms). Computing in two ways the sum of all digits above, we obtain $n-u=v-1$ and $u=n-v+1$. Then, the first $n-v$ terms in the $v$ th block are zeroes, and the next $v$ terms are ones, due to the sum of all terms in this block. The statement is proved.

We are left to check that the sequence obtained satisfies the condition. Notice that $a_{i} \leq a_{i+n}$ for all $1 \leq i \leq n^{2}$. Moreover, if $1 \leq u \leq n$ and $0 \leq v \leq n-1$, then $a_{u+v n}<a_{u+v n+n}$ exactly when $u+v=n$. In this case we have $u+v n=n+v(n-1)$.

Consider now an arbitrary index $0 \leq i \leq n^{2}-n$. Clearly, there exists an integer $v$ such that $n+v(n-1) \in[i+1, i+n]$. Then, applying the above inequalities we obtain that condition (b) is valid.
Solution 2. Similarly to Solution 1, we introduce the notation $S(k, l]$ and obtain (2), (3), and (4) in the same way. The sum of all elements of the sequence can be computed as

$$
S\left(0, n^{2}+n\right]=S(0, n]+S(n, 2 n]+\ldots+S\left(n^{2}, n^{2}+n\right]=0+1+\ldots+n
$$

For an arbitrary integer $0 \leq u \leq n$, consider the numbers

$$
\begin{equation*}
S(u, u+n]<S(u+n, u+2 n]<\ldots<S\left(u+(n-1) n, u+n^{2}\right] . \tag{5}
\end{equation*}
$$

They are $n$ distinct integers from the $n+1$ possible values $0,1,2, \ldots, n$. Denote by $m$ the "missing" value which is not listed. We determine $m$ from $S\left(0, n^{2}+n\right]$. Write this sum as
$S\left(0, n^{2}+n\right]=S(0, u]+S(u, u+n]+S(u+n, u+2 n]+\ldots+S\left(u+(n-1) n, u+n^{2}\right]+S\left(u+n^{2}, n^{2}+n\right]$.
Since $a_{1}=a_{2}=\ldots=a_{u}=0$ and $a_{u+n^{2}+1}=\ldots=a_{n^{2}+n}=1$, we have $S(0, u]=0$ and $S\left(u+n^{2}, n+n^{2}\right]=n-u$. Then

$$
0+1+\ldots+n=S\left(0, n^{2}+n\right]=0+((0+1+\ldots+n)-m)+(n-u)
$$

so $m=n-u$.
Hence, the numbers listed in (5) are $0,1, \ldots, n-u-1$ and $n-u+1, \ldots, n$, respectively, therefore

$$
S(u+v n, u+(v+1) n]=\left\{\begin{array}{ll}
v, & v \leq n-u-1,  \tag{6}\\
v+1, & v \geq n-u
\end{array} \quad \text { for all } 0 \leq u \leq n, 0 \leq v \leq n-1\right.
$$

Conditions (6), together with (3), provide a system of linear equations in variables $a_{i}$. Now we solve this system and show that the solution is unique and satisfies conditions (a) and (b).

First, observe that any solution of the system (3), (6) satisfies the condition (b). By the construction, equations (6) immediately imply (5). On the other hand, all inequalities mentioned in condition (b) are included into the chain (5) for some value of $u$.

Next, note that the system (3), (6) is redundant. The numbers $S(k n,(k+1) n]$, where $1 \leq k \leq n-1$, appear twice in (6). For $u=0$ and $v=k$ we have $v \leq n-u-1$, and (6) gives $S(k n,(k+1) n]=v=k$. For $u=n$ and $v=k-1$ we have $v \geq n-u$ and we obtain the same value, $S(k n,(k+1) n]=v+1=k$. Therefore, deleting one equation from each redundant pair, we can make every sum $S(k, k+n$ ] appear exactly once on the left-hand side in (6).

Now, from (3), (6), the sequence $\left(a_{i}\right)$ can be reconstructed inductively by
$a_{1}=a_{2}=\ldots=a_{n-1}=0, \quad a_{k+n}=S(k, k+n]-\left(a_{k+1}+a_{k+2}+\ldots+a_{k+n-1}\right) \quad\left(0 \leq k \leq n^{2}\right)$,
taking the values of $S(k, k+n]$ from (6). This means first that there exists at most one solution of our system. Conversely, the constructed sequence obviously satisfies all equations (3), (6) (the only missing equation is $a_{n}=0$, which follows from $S(0, n]=0$ ). Hence it satisfies condition (b), and we are left to check condition (a) only.

For arbitrary integers $1 \leq u, t \leq n$ we get

$$
\begin{aligned}
a_{u+t n}-a_{u+(t-1) n} & =S(u+(t-1) n, u+t n]-S((u-1)+(t-1) n,(u-1)+t n] \\
& = \begin{cases}(t-1)-(t-1)=0, & t \leq n-u \\
t-(t-1)=1, & t=n-u+1 \\
t-t=0, & t \geq n-u+2\end{cases}
\end{aligned}
$$

Since $a_{u}=0$, we have

$$
a_{u+v n}=a_{u+v n}-a_{u}=\sum_{t=1}^{v}\left(a_{u+t n}-a_{u+(t-1) n}\right)
$$

for all $1 \leq u, v \leq n$. If $v<n-u+1$ then all terms are 0 on the right-hand side. If $v \geq n-u+1$, then variable $t$ attains the value $n-u+1$ once. Hence,

$$
a_{u+v n}= \begin{cases}0, & u+v \leq n \\ 1, & u+v \geq n+1\end{cases}
$$

according with (1). Note that the formula is valid for $v=0$ as well.
Finally, we presented the direct formula for $\left(a_{i}\right)$, and we have proved that it satisfies condition (a). So, the solution is complete.

C2. A unit square is dissected into $n>1$ rectangles such that their sides are parallel to the sides of the square. Any line, parallel to a side of the square and intersecting its interior, also intersects the interior of some rectangle. Prove that in this dissection, there exists a rectangle having no point on the boundary of the square.
(Japan)
Solution 1. Call the directions of the sides of the square horizontal and vertical. A horizontal or vertical line, which intersects the interior of the square but does not intersect the interior of any rectangle, will be called a splitting line. A rectangle having no point on the boundary of the square will be called an interior rectangle.

Suppose, to the contrary, that there exists a dissection of the square into more than one rectangle, such that no interior rectangle and no splitting line appear. Consider such a dissection with the least possible number of rectangles. Notice that this number of rectangles is greater than 2 , otherwise their common side provides a splitting line.

If there exist two rectangles having a common side, then we can replace them by their union (see Figure 1). The number of rectangles was greater than 2, so in a new dissection it is greater than 1. Clearly, in the new dissection, there is also no splitting line as well as no interior rectangle. This contradicts the choice of the original dissection.

Denote the initial square by $A B C D$, with $A$ and $B$ being respectively the lower left and lower right vertices. Consider those two rectangles $a$ and $b$ containing vertices $A$ and $B$, respectively. (Note that $a \neq b$, otherwise its top side provides a splitting line.) We can assume that the height of $a$ is not greater than that of $b$. Then consider the rectangle $c$ neighboring to the lower right corner of $a$ (it may happen that $c=b$ ). By aforementioned, the heights of $a$ and $c$ are distinct. Then two cases are possible.


Figure 1


Figure 2


Figure 3

Case 1. The height of $c$ is less than that of $a$. Consider the rectangle $d$ which is adjacent to both $a$ and $c$, i.e. the one containing the angle marked in Figure 2. This rectangle has no common point with $B C$ (since $a$ is not higher than $b$ ), as well as no common point with $A B$ or with $A D$ (obviously). Then $d$ has a common point with $C D$, and its left side provides a splitting line. Contradiction.

Case 2. The height of $c$ is greater than that of $a$. Analogously, consider the rectangle $d$ containing the angle marked on Figure 3. It has no common point with $A D$ (otherwise it has a common side with $a$ ), as well as no common point with $A B$ or with $B C$ (obviously). Then $d$ has a common point with $C D$. Hence its right side provides a splitting line, and we get the contradiction again.

Solution 2. Again, we suppose the contrary. Consider an arbitrary counterexample. Then we know that each rectangle is attached to at least one side of the square. Observe that a rectangle cannot be attached to two opposite sides, otherwise one of its sides lies on a splitting line.

We say that two rectangles are opposite if they are attached to opposite sides of $A B C D$. We claim that there exist two opposite rectangles having a common point.

Consider the union $L$ of all rectangles attached to the left. Assume, to the contrary, that $L$ has no common point with the rectangles attached to the right. Take a polygonal line $p$ connecting the top and the bottom sides of the square and passing close from the right to the boundary of $L$ (see Figure 4). Then all its points belong to the rectangles attached either to the top or to the bottom. Moreover, the upper end-point of $p$ belongs to a rectangle attached to the top, and the lower one belongs to an other rectangle attached to the bottom. Hence, there is a point on $p$ where some rectangles attached to the top and to the bottom meet each other. So, there always exists a pair of neighboring opposite rectangles.


Now, take two opposite neighboring rectangles $a$ and $b$. We can assume that $a$ is attached to the left and $b$ is attached to the right. Let $X$ be their common point. If $X$ belongs to their horizontal sides (in particular, $X$ may appear to be a common vertex of $a$ and $b$ ), then these sides provide a splitting line (see Figure 5). Otherwise, $X$ lies on the vertical sides. Let $\ell$ be the line containing these sides.

Since $\ell$ is not a splitting line, it intersects the interior of some rectangle. Let $c$ be such a rectangle, closest to $X$; we can assume that $c$ lies above $X$. Let $Y$ be the common point of $\ell$ and the bottom side of $c$ (see Figure 6). Then $Y$ is also a vertex of two rectangles lying below $c$.

So, let $Y$ be the upper-right and upper-left corners of the rectangles $a^{\prime}$ and $b^{\prime}$, respectively. Then $a^{\prime}$ and $b^{\prime}$ are situated not lower than $a$ and $b$, respectively (it may happen that $a=a^{\prime}$ or $b=b^{\prime}$ ). We claim that $a^{\prime}$ is attached to the left. If $a=a^{\prime}$ then of course it is. If $a \neq a^{\prime}$ then $a^{\prime}$ is above $a$, below $c$ and to the left from $b^{\prime}$. Hence, it can be attached to the left only.

Analogously, $b^{\prime}$ is attached to the right. Now, the top sides of these two rectangles pass through $Y$, hence they provide a splitting line again. This last contradiction completes the proof.

C3. Find all positive integers $n$, for which the numbers in the set $S=\{1,2, \ldots, n\}$ can be colored red and blue, with the following condition being satisfied: the set $S \times S \times S$ contains exactly 2007 ordered triples $(x, y, z)$ such that (i) $x, y, z$ are of the same color and (ii) $x+y+z$ is divisible by $n$.
(Netherlands)
Answer. $n=69$ and $n=84$.
Solution. Suppose that the numbers $1,2, \ldots, n$ are colored red and blue. Denote by $R$ and $B$ the sets of red and blue numbers, respectively; let $|R|=r$ and $|B|=b=n-r$. Call a triple $(x, y, z) \in S \times S \times S$ monochromatic if $x, y, z$ have the same color, and bichromatic otherwise. Call a triple $(x, y, z)$ divisible if $x+y+z$ is divisible by $n$. We claim that there are exactly $r^{2}-r b+b^{2}$ divisible monochromatic triples.

For any pair $(x, y) \in S \times S$ there exists a unique $z_{x, y} \in S$ such that the triple $\left(x, y, z_{x, y}\right)$ is divisible; so there are exactly $n^{2}$ divisible triples. Furthermore, if a divisible triple $(x, y, z)$ is bichromatic, then among $x, y, z$ there are either one blue and two red numbers, or vice versa. In both cases, exactly one of the pairs $(x, y),(y, z)$ and $(z, x)$ belongs to the set $R \times B$. Assign such pair to the triple $(x, y, z)$.

Conversely, consider any pair $(x, y) \in R \times B$, and denote $z=z_{x, y}$. Since $x \neq y$, the triples $(x, y, z),(y, z, x)$ and $(z, x, y)$ are distinct, and $(x, y)$ is assigned to each of them. On the other hand, if $(x, y)$ is assigned to some triple, then this triple is clearly one of those mentioned above. So each pair in $R \times B$ is assigned exactly three times.

Thus, the number of bichromatic divisible triples is three times the number of elements in $R \times B$, and the number of monochromatic ones is $n^{2}-3 r b=(r+b)^{2}-3 r b=r^{2}-r b+b^{2}$, as claimed.

So, to find all values of $n$ for which the desired coloring is possible, we have to find all $n$, for which there exists a decomposition $n=r+b$ with $r^{2}-r b+b^{2}=2007$. Therefore, $9 \mid r^{2}-r b+b^{2}=(r+b)^{2}-3 r b$. From this it consequently follows that $3|r+b, 3| r b$, and then $3|r, 3| b$. Set $r=3 s, b=3 c$. We can assume that $s \geq c$. We have $s^{2}-s c+c^{2}=223$.

Furthermore,

$$
892=4\left(s^{2}-s c+c^{2}\right)=(2 c-s)^{2}+3 s^{2} \geq 3 s^{2} \geq 3 s^{2}-3 c(s-c)=3\left(s^{2}-s c+c^{2}\right)=669
$$

so $297 \geq s^{2} \geq 223$ and $17 \geq s \geq 15$. If $s=15$ then

$$
c(15-c)=c(s-c)=s^{2}-\left(s^{2}-s c+c^{2}\right)=15^{2}-223=2
$$

which is impossible for an integer $c$. In a similar way, if $s=16$ then $c(16-c)=33$, which is also impossible. Finally, if $s=17$ then $c(17-c)=66$, and the solutions are $c=6$ and $c=11$. Hence, $(r, b)=(51,18)$ or $(r, b)=(51,33)$, and the possible values of $n$ are $n=51+18=69$ and $n=51+33=84$.
Comment. After the formula for the number of monochromatic divisible triples is found, the solution can be finished in various ways. The one presented is aimed to decrease the number of considered cases.
$\mathbf{C 4}$. Let $A_{0}=\left(a_{1}, \ldots, a_{n}\right)$ be a finite sequence of real numbers. For each $k \geq 0$, from the sequence $A_{k}=\left(x_{1}, \ldots, x_{n}\right)$ we construct a new sequence $A_{k+1}$ in the following way.

1. We choose a partition $\{1, \ldots, n\}=I \cup J$, where $I$ and $J$ are two disjoint sets, such that the expression

$$
\left|\sum_{i \in I} x_{i}-\sum_{j \in J} x_{j}\right|
$$

attains the smallest possible value. (We allow the sets $I$ or $J$ to be empty; in this case the corresponding sum is 0 .) If there are several such partitions, one is chosen arbitrarily.
2. We set $A_{k+1}=\left(y_{1}, \ldots, y_{n}\right)$, where $y_{i}=x_{i}+1$ if $i \in I$, and $y_{i}=x_{i}-1$ if $i \in J$.

Prove that for some $k$, the sequence $A_{k}$ contains an element $x$ such that $|x| \geq n / 2$.
(Iran)

## Solution.

Lemma. Suppose that all terms of the sequence $\left(x_{1}, \ldots, x_{n}\right)$ satisfy the inequality $\left|x_{i}\right|<a$. Then there exists a partition $\{1,2, \ldots, n\}=I \cup J$ into two disjoint sets such that

$$
\begin{equation*}
\left|\sum_{i \in I} x_{i}-\sum_{j \in J} x_{j}\right|<a \tag{1}
\end{equation*}
$$

Proof. Apply an induction on $n$. The base case $n=1$ is trivial. For the induction step, consider a sequence $\left(x_{1}, \ldots, x_{n}\right)(n>1)$. By the induction hypothesis there exists a splitting $\{1, \ldots, n-1\}=I^{\prime} \cup J^{\prime}$ such that

$$
\left|\sum_{i \in I^{\prime}} x_{i}-\sum_{j \in J^{\prime}} x_{j}\right|<a
$$

For convenience, suppose that $\sum_{i \in I^{\prime}} x_{i} \geq \sum_{j \in J^{\prime}} x_{j}$. If $x_{n} \geq 0$ then choose $I=I^{\prime}, J=J \cup\{n\}$; otherwise choose $I=I^{\prime} \cup\{n\}, J=J^{\prime}$. In both cases, we have $\sum_{i \in I^{\prime}} x_{i}-\sum_{j \in J^{\prime}} x_{j} \in[0, a)$ and $\left|x_{n}\right| \in[0, a)$; hence

$$
\sum_{i \in I} x_{i}-\sum_{j \in J} x_{j}=\sum_{i \in I^{\prime}} x_{i}-\sum_{j \in J^{\prime}} x_{j}-\left|x_{n}\right| \in(-a, a),
$$

as desired.
Let us turn now to the problem. To the contrary, assume that for all $k$, all the numbers in $A_{k}$ lie in interval $(-n / 2, n / 2)$. Consider an arbitrary sequence $A_{k}=\left(b_{1}, \ldots, b_{n}\right)$. To obtain the term $b_{i}$, we increased and decreased number $a_{i}$ by one several times. Therefore $b_{i}-a_{i}$ is always an integer, and there are not more than $n$ possible values for $b_{i}$. So, there are not more than $n^{n}$ distinct possible sequences $A_{k}$, and hence two of the sequences $A_{1}, A_{2}, \ldots, A_{n^{n}+1}$ should be identical, say $A_{p}=A_{q}$ for some $p<q$.

For any positive integer $k$, let $S_{k}$ be the sum of squares of elements in $A_{k}$. Consider two consecutive sequences $A_{k}=\left(x_{1}, \ldots, x_{n}\right)$ and $A_{k+1}=\left(y_{1}, \ldots, y_{n}\right)$. Let $\{1,2, \ldots, n\}=I \cup J$ be the partition used in this step - that is, $y_{i}=x_{i}+1$ for all $i \in I$ and $y_{j}=x_{j}-1$ for all $j \in J$. Since the value of $\left|\sum_{i \in I} x_{i}-\sum_{j \in J} x_{j}\right|$ is the smallest possible, the Lemma implies that it is less than $n / 2$. Then we have
$S_{k+1}-S_{k}=\sum_{i \in I}\left(\left(x_{i}+1\right)^{2}-x_{i}^{2}\right)+\sum_{j \in J}\left(\left(x_{j}-1\right)^{2}-x_{j}^{2}\right)=n+2\left(\sum_{i \in I} x_{i}-\sum_{j \in J} x_{j}\right)>n-2 \cdot \frac{n}{2}=0$.
Thus we obtain $S_{q}>S_{q-1}>\cdots>S_{p}$. This is impossible since $A_{p}=A_{q}$ and hence $S_{p}=S_{q}$.

C5. In the Cartesian coordinate plane define the strip $S_{n}=\{(x, y) \mid n \leq x<n+1\}$ for every integer $n$. Assume that each strip $S_{n}$ is colored either red or blue, and let $a$ and $b$ be two distinct positive integers. Prove that there exists a rectangle with side lengths $a$ and $b$ such that its vertices have the same color.
(Romania)
Solution. If $S_{n}$ and $S_{n+a}$ have the same color for some integer $n$, then we can choose the rectangle with vertices $(n, 0) \in S_{n},(n, b) \in S_{n},(n+a, 0) \in S_{n+a}$, and $(n+a, b) \in S_{n+a}$, and we are done. So it can be assumed that $S_{n}$ and $S_{n+a}$ have opposite colors for each $n$.

Similarly, it also can be assumed that $S_{n}$ and $S_{n+b}$ have opposite colors. Then, by induction on $|p|+|q|$, we obtain that for arbitrary integers $p$ and $q$, strips $S_{n}$ and $S_{n+p a+q b}$ have the same color if $p+q$ is even, and these two strips have opposite colors if $p+q$ is odd.

Let $d=\operatorname{gcd}(a, b), a_{1}=a / d$ and $b_{1}=b / d$. Apply the result above for $p=b_{1}$ and $q=-a_{1}$. The strips $S_{0}$ and $S_{0+b_{1} a-a_{1} b}$ are identical and therefore they have the same color. Hence, $a_{1}+b_{1}$ is even. By the construction, $a_{1}$ and $b_{1}$ are coprime, so this is possible only if both are odd.

Without loss of generality, we can assume $a>b$. Then $a_{1}>b_{1} \geq 1$, so $a_{1} \geq 3$.
Choose integers $k$ and $\ell$ such that $k a_{1}-\ell b_{1}=1$ and therefore $k a-\ell b=d$. Since $a_{1}$ and $b_{1}$ are odd, $k+\ell$ is odd as well. Hence, for every integer $n$, strips $S_{n}$ and $S_{n+k a-\ell b}=S_{n+d}$ have opposite colors. This also implies that the coloring is periodic with period 2d, i.e. strips $S_{n}$ and $S_{n+2 d}$ have the same color for every $n$.


Figure 1
We will construct the desired rectangle $A B C D$ with $A B=C D=a$ and $B C=A D=b$ in a position such that vertex $A$ lies on the $x$-axis, and the projection of side $A B$ onto the $x$-axis is of length $2 d$ (see Figure 1). This is possible since $a=a_{1} d>2 d$. The coordinates of the vertices will have the forms

$$
A=(t, 0), \quad B=\left(t+2 d, y_{1}\right), \quad C=\left(u+2 d, y_{2}\right), \quad D=\left(u, y_{3}\right) .
$$

Let $\varphi=\sqrt{a_{1}^{2}-4}$. By Pythagoras' theorem,

$$
y_{1}=B B_{0}=\sqrt{a^{2}-4 d^{2}}=d \sqrt{a_{1}^{2}-4}=d \varphi .
$$

So, by the similar triangles $A D D_{0}$ and $B A B_{0}$, we have the constraint

$$
\begin{equation*}
u-t=A D_{0}=\frac{A D}{A B} \cdot B B_{0}=\frac{b d}{a} \varphi \tag{1}
\end{equation*}
$$

for numbers $t$ and $u$. Computing the numbers $y_{2}$ and $y_{3}$ is not required since they have no effect to the colors.

Observe that the number $\varphi$ is irrational, because $\varphi^{2}$ is an integer, but $\varphi$ is not: $a_{1}>\varphi \geq$ $\sqrt{a_{1}^{2}-2 a_{1}+2}>a_{1}-1$.

By the periodicity, points $A$ and $B$ have the same color; similarly, points $C$ and $D$ have the same color. Furthermore, these colors depend only on the values of $t$ and $u$. So it is sufficient to choose numbers $t$ and $u$ such that vertices $A$ and $D$ have the same color.

Let $w$ be the largest positive integer such that there exist $w$ consecutive strips $S_{n_{0}}, S_{n_{0}+1}, \ldots$, $S_{n_{0}+w-1}$ with the same color, say red. (Since $S_{n_{0}+d}$ must be blue, we have $w \leq d$.) We will choose $t$ from the interval $\left(n_{0}, n_{0}+w\right)$.


Figure 2
Consider the interval $I=\left(n_{0}+\frac{b d}{a} \varphi, n_{0}+\frac{b d}{a} \varphi+w\right)$ on the $x$-axis (see Figure 2). Its length is $w$, and the end-points are irrational. Therefore, this interval intersects $w+1$ consecutive strips. Since at most $w$ consecutive strips may have the same color, interval $I$ must contain both red and blue points. Choose $u \in I$ such that the line $x=u$ is red and set $t=u-\frac{b d}{a} \varphi$, according to the constraint (1). Then $t \in\left(n_{0}, n_{0}+w\right)$ and $A=(t, 0)$ is red as well as $D=\left(u, y_{3}\right)$.

Hence, variables $u$ and $t$ can be set such that they provide a rectangle with four red vertices.
Comment. The statement is false for squares, i.e. in the case $a=b$. If strips $S_{2 k a}, S_{2 k a+1}, \ldots$, $S_{(2 k+1) a-1}$ are red, and strips $S_{(2 k+1) a}, S_{(2 k+1) a+1}, \ldots, S_{(2 k+2) a-1}$ are blue for every integer $k$, then each square of size $a \times a$ has at least one red and at least one blue vertex as well.

C6. In a mathematical competition some competitors are friends; friendship is always mutual. Call a group of competitors a clique if each two of them are friends. The number of members in a clique is called its size.

It is known that the largest size of cliques is even. Prove that the competitors can be arranged in two rooms such that the largest size of cliques in one room is the same as the largest size of cliques in the other room.
(Russia)
Solution. We present an algorithm to arrange the competitors. Let the two rooms be Room $A$ and Room B. We start with an initial arrangement, and then we modify it several times by sending one person to the other room. At any state of the algorithm, $A$ and $B$ denote the sets of the competitors in the rooms, and $c(A)$ and $c(B)$ denote the largest sizes of cliques in the rooms, respectively.
Step 1. Let $M$ be one of the cliques of largest size, $|M|=2 m$. Send all members of $M$ to Room $A$ and all other competitors to Room B.

Since $M$ is a clique of the largest size, we have $c(A)=|M| \geq c(B)$.
Step 2. While $c(A)>c(B)$, send one person from Room $A$ to Room $B$.


Note that $c(A)>c(B)$ implies that Room $A$ is not empty.
In each step, $c(A)$ decreases by one and $c(B)$ increases by at most one. So at the end we have $c(A) \leq c(B) \leq c(A)+1$.

We also have $c(A)=|A| \geq m$ at the end. Otherwise we would have at least $m+1$ members of $M$ in Room $B$ and at most $m-1$ in Room $A$, implying $c(B)-c(A) \geq(m+1)-(m-1)=2$.
Step 3. Let $k=c(A)$. If $c(B)=k$ then STOP.
If we reached $c(A)=c(B)=k$ then we have found the desired arrangement.
In all other cases we have $c(B)=k+1$.
From the estimate above we also know that $k=|A|=|A \cap M| \geq m$ and $|B \cap M| \leq m$.
Step 4. If there exists a competitor $x \in B \cap M$ and a clique $C \subset B$ such that $|C|=k+1$ and $x \notin C$, then move $x$ to Room $A$ and STOP.


After moving $x$ back to Room $A$, we will have $k+1$ members of $M$ in Room $A$, thus $c(A)=k+1$. Due to $x \notin C, c(B)=|C|$ is not decreased, and after this step we have $c(A)=c(B)=k+1$.

If there is no such competitor $x$, then in Room $B$, all cliques of size $k+1$ contain $B \cap M$ as a subset.
Step 5. While $c(B)=k+1$, choose a clique $C \subset B$ such that $|C|=k+1$ and move one member of $C \backslash M$ to Room $A$.


Note that $|C|=k+1>m \geq|B \cap M|$, so $C \backslash M$ cannot be empty.
Every time we move a single person from Room $B$ to Room $A$, so $c(B)$ decreases by at most 1. Hence, at the end of this loop we have $c(B)=k$.

In Room $A$ we have the clique $A \cap M$ with size $|A \cap M|=k$ thus $c(A) \geq k$. We prove that there is no clique of larger size there. Let $Q \subset A$ be an arbitrary clique. We show that $|Q| \leq k$.


In Room $A$, and specially in set $Q$, there can be two types of competitors:

- Some members of $M$. Since $M$ is a clique, they are friends with all members of $B \cap M$.
- Competitors which were moved to Room $A$ in Step 5. Each of them has been in a clique with $B \cap M$ so they are also friends with all members of $B \cap M$.

Hence, all members of $Q$ are friends with all members of $B \cap M$. Sets $Q$ and $B \cap M$ are cliques themselves, so $Q \cup(B \cap M)$ is also a clique. Since $M$ is a clique of the largest size,

$$
|M| \geq|Q \cup(B \cap M)|=|Q|+|B \cap M|=|Q|+|M|-|A \cap M|,
$$

therefore

$$
|Q| \leq|A \cap M|=k
$$

Finally, after Step 5 we have $c(A)=c(B)=k$.
Comment. Obviously, the statement is false without the assumption that the largest clique size is even.

C7. Let $\alpha<\frac{3-\sqrt{5}}{2}$ be a positive real number. Prove that there exist positive integers $n$ and $p>\alpha \cdot 2^{n}$ for which one can select $2 p$ pairwise distinct subsets $S_{1}, \ldots, S_{p}, T_{1}, \ldots, T_{p}$ of the set $\{1,2, \ldots, n\}$ such that $S_{i} \cap T_{j} \neq \varnothing$ for all $1 \leq i, j \leq p$.

Solution. Let $k$ and $m$ be positive integers (to be determined later) and set $n=k m$. Decompose the set $\{1,2, \ldots, n\}$ into $k$ disjoint subsets, each of size $m$; denote these subsets by $A_{1}, \ldots, A_{k}$. Define the following families of sets:

$$
\begin{aligned}
\mathcal{S} & =\left\{S \subset\{1,2, \ldots, n\}: \forall i S \cap A_{i} \neq \varnothing\right\} \\
\mathcal{T}_{1} & =\left\{T \subset\{1,2, \ldots, n\}: \exists i A_{i} \subset T\right\}, \quad \mathcal{T}=\mathcal{T}_{1} \backslash \mathcal{S}
\end{aligned}
$$

For each set $T \in \mathcal{T} \subset \mathcal{T}_{1}$, there exists an index $1 \leq i \leq k$ such that $A_{i} \subset T$. Then for all $S \in \mathcal{S}$, $S \cap T \supset S \cap A_{i} \neq \varnothing$. Hence, each $S \in \mathcal{S}$ and each $T \in \mathcal{T}$ have at least one common element.

Below we show that the numbers $m$ and $k$ can be chosen such that $|\mathcal{S}|,|\mathcal{T}|>\alpha \cdot 2^{n}$. Then, choosing $p=\min \{|\mathcal{S}|,|\mathcal{T}|\}$, one can select the desired $2 p$ sets $S_{1}, \ldots, S_{p}$ and $T_{1}, \ldots, T_{p}$ from families $\mathcal{S}$ and $\mathcal{T}$, respectively. Since families $\mathcal{S}$ and $\mathcal{T}$ are disjoint, sets $S_{i}$ and $T_{j}$ will be pairwise distinct.

To count the sets $S \in \mathcal{S}$, observe that each $A_{i}$ has $2^{m}-1$ nonempty subsets so we have $2^{m}-1$ choices for $S \cap A_{i}$. These intersections uniquely determine set $S$, so

$$
\begin{equation*}
|\mathcal{S}|=\left(2^{m}-1\right)^{k} \tag{1}
\end{equation*}
$$

Similarly, if a set $H \subset\{1,2, \ldots, n\}$ does not contain a certain set $A_{i}$ then we have $2^{m}-1$ choices for $H \cap A_{i}$ : all subsets of $A_{i}$, except $A_{i}$ itself. Therefore, the complement of $\mathcal{T}_{1}$ contains $\left(2^{m}-1\right)^{k}$ sets and

$$
\begin{equation*}
\left|\mathcal{T}_{1}\right|=2^{k m}-\left(2^{m}-1\right)^{k} . \tag{2}
\end{equation*}
$$

Next consider the family $\mathcal{S} \backslash \mathcal{T}_{1}$. If a set $S$ intersects all $A_{i}$ but does not contain any of them, then there exists $2^{m}-2$ possible values for each $S \cap A_{i}$ : all subsets of $A_{i}$ except $\varnothing$ and $A_{i}$. Therefore the number of such sets $S$ is $\left(2^{m}-2\right)^{k}$, so

$$
\begin{equation*}
\left|\mathcal{S} \backslash \mathcal{T}_{1}\right|=\left(2^{m}-2\right)^{k} \tag{3}
\end{equation*}
$$

From (1), (2), and (3) we obtain

$$
|\mathcal{T}|=\left|\mathcal{T}_{1}\right|-\left|\mathcal{S} \cap \mathcal{T}_{1}\right|=\left|\mathcal{T}_{1}\right|-\left(|\mathcal{S}|-\left|\mathcal{S} \backslash \mathcal{T}_{1}\right|\right)=2^{k m}-2\left(2^{m}-1\right)^{k}+\left(2^{m}-2\right)^{k}
$$

Let $\delta=\frac{3-\sqrt{5}}{2}$ and $k=k(m)=\left[2^{m} \log \frac{1}{\delta}\right]$. Then

$$
\lim _{m \rightarrow \infty} \frac{|\mathcal{S}|}{2^{k m}}=\lim _{m \rightarrow \infty}\left(1-\frac{1}{2^{m}}\right)^{k}=\exp \left(-\lim _{m \rightarrow \infty} \frac{k}{2^{m}}\right)=\delta
$$

and similarly

$$
\lim _{m \rightarrow \infty} \frac{|\mathcal{T}|}{2^{k m}}=1-2 \lim _{m \rightarrow \infty}\left(1-\frac{1}{2^{m}}\right)^{k}+\lim _{m \rightarrow \infty}\left(1-\frac{2}{2^{m}}\right)^{k}=1-2 \delta+\delta^{2}=\delta
$$

Hence, if $m$ is sufficiently large then $\frac{|\mathcal{S}|}{2^{m k}}$ and $\frac{|\mathcal{T}|}{2^{m k}}$ are greater than $\alpha$ (since $\alpha<\delta$ ). So $|\mathcal{S}|,|\mathcal{T}|>\alpha \cdot 2^{m k}=\alpha \cdot 2^{n}$.
Comment. It can be proved that the constant $\frac{3-\sqrt{5}}{2}$ is sharp. Actually, if $S_{1}, \ldots, S_{p}, T_{1}, \ldots, T_{p}$ are distinct subsets of $\{1,2, \ldots, n\}$ such that each $S_{i}$ intersects each $T_{j}$, then $p<\frac{3-\sqrt{5}}{2} \cdot 2^{n}$.

C8. Given a convex $n$-gon $P$ in the plane. For every three vertices of $P$, consider the triangle determined by them. Call such a triangle good if all its sides are of unit length.

Prove that there are not more than $\frac{2}{3} n$ good triangles.
(Ukraine)
Solution. Consider all good triangles containing a certain vertex $A$. The other two vertices of any such triangle lie on the circle $\omega_{A}$ with unit radius and center $A$. Since $P$ is convex, all these vertices lie on an arc of angle less than $180^{\circ}$. Let $L_{A} R_{A}$ be the shortest such arc, oriented clockwise (see Figure 1). Each of segments $A L_{A}$ and $A R_{A}$ belongs to a unique good triangle. We say that the good triangle with side $A L_{A}$ is assigned counterclockwise to $A$, and the second one, with side $A R_{A}$, is assigned clockwise to $A$. In those cases when there is a single good triangle containing vertex $A$, this triangle is assigned to $A$ twice.

There are at most two assignments to each vertex of the polygon. (Vertices which do not belong to any good triangle have no assignment.) So the number of assignments is at most $2 n$.

Consider an arbitrary good triangle $A B C$, with vertices arranged clockwise. We prove that $A B C$ is assigned to its vertices at least three times. Then, denoting the number of good triangles by $t$, we obtain that the number $K$ of all assignments is at most $2 n$, while it is not less than $3 t$. Then $3 t \leq K \leq 2 n$, as required.

Actually, we prove that triangle $A B C$ is assigned either counterclockwise to $C$ or clockwise to $B$. Then, by the cyclic symmetry of the vertices, we obtain that triangle $A B C$ is assigned either counterclockwise to $A$ or clockwise to $C$, and either counterclockwise to $B$ or clockwise to $A$, providing the claim.


Figure 1


Figure 2

Assume, to the contrary, that $L_{C} \neq A$ and $R_{B} \neq A$. Denote by $A^{\prime}, B^{\prime}, C^{\prime}$ the intersection points of circles $\omega_{A}, \omega_{B}$ and $\omega_{C}$, distinct from $A, B, C$ (see Figure 2). Let $C L_{C} L_{C}^{\prime}$ be the good triangle containing $C L_{C}$. Observe that the angle of arc $L_{C} A$ is less than $120^{\circ}$. Then one of the points $L_{C}$ and $L_{C}^{\prime}$ belongs to arc $B^{\prime} A$ of $\omega_{C}$; let this point be $X$. In the case when $L_{C}=B^{\prime}$ and $L_{C}^{\prime}=A$, choose $X=B^{\prime}$.

Analogously, considering the good triangle $B R_{B}^{\prime} R_{B}$ which contains $B R_{B}$ as an edge, we see that one of the points $R_{B}$ and $R_{B}^{\prime}$ lies on arc $A C^{\prime}$ of $\omega_{B}$. Denote this point by $Y, Y \neq A$. Then angles $X A Y, Y A B, B A C$ and $C A X$ (oriented clockwise) are not greater than $180^{\circ}$. Hence, point $A$ lies in quadrilateral $X Y B C$ (either in its interior or on segment $X Y$ ). This is impossible, since all these five points are vertices of $P$.

Hence, each good triangle has at least three assignments, and the statement is proved.
Comment 1. Considering a diameter $A B$ of the polygon, one can prove that every good triangle containing either $A$ or $B$ has at least four assignments. This observation leads to $t \leq\left\lfloor\frac{2}{3}(n-1)\right\rfloor$.

Comment 2. The result $t \leq\left\lfloor\frac{2}{3}(n-1)\right\rfloor$ is sharp. To construct a polygon with $n=3 k+1$ vertices and $t=2 k$ triangles, take a rhombus $A B_{1} C_{1} D_{1}$ with unit side length and $\angle B_{1}=60^{\circ}$. Then rotate it around $A$ by small angles obtaining rhombi $A B_{2} C_{2} D_{2}, \ldots, A B_{k} C_{k} D_{k}$ (see Figure 3). The polygon $A B_{1} \ldots B_{k} C_{1} \ldots C_{k} D_{1} \ldots D_{k}$ has $3 k+1$ vertices and contains $2 k$ good triangles.

The construction for $n=3 k$ and $n=3 k-1$ can be obtained by deleting vertices $D_{n}$ and $D_{n-1}$.


Figure 3

## Geometry

G1. In triangle $A B C$, the angle bisector at vertex $C$ intersects the circumcircle and the perpendicular bisectors of sides $B C$ and $C A$ at points $R, P$, and $Q$, respectively. The midpoints of $B C$ and $C A$ are $S$ and $T$, respectively. Prove that triangles $R Q T$ and $R P S$ have the same area.
(Czech Republic)
Solution 1. If $A C=B C$ then triangle $A B C$ is isosceles, triangles $R Q T$ and $R P S$ are symmetric about the bisector $C R$ and the statement is trivial. If $A C \neq B C$ then it can be assumed without loss of generality that $A C<B C$.


Denote the circumcenter by $O$. The right triangles $C T Q$ and $C S P$ have equal angles at vertex $C$, so they are similar, $\angle C P S=\angle C Q T=\angle O Q P$ and

$$
\begin{equation*}
\frac{Q T}{P S}=\frac{C Q}{C P} \tag{1}
\end{equation*}
$$

Let $\ell$ be the perpendicular bisector of chord $C R$; of course, $\ell$ passes through the circumcenter $O$. Due to the equal angles at $P$ and $Q$, triangle $O P Q$ is isosceles with $O P=O Q$. Then line $\ell$ is the axis of symmetry in this triangle as well. Therefore, points $P$ and $Q$ lie symmetrically on line segment $C R$,

$$
\begin{equation*}
R P=C Q \quad \text { and } \quad R Q=C P \tag{2}
\end{equation*}
$$

Triangles $R Q T$ and $R P S$ have equal angles at vertices $Q$ and $P$, respectively. Then

$$
\frac{\operatorname{area}(R Q T)}{\operatorname{area}(R P S)}=\frac{\frac{1}{2} \cdot R Q \cdot Q T \cdot \sin \angle R Q T}{\frac{1}{2} \cdot R P \cdot P S \cdot \sin \angle R P S}=\frac{R Q}{R P} \cdot \frac{Q T}{P S}
$$

Substituting (1) and (2),

$$
\frac{\operatorname{area}(R Q T)}{\operatorname{area}(R P S)}=\frac{R Q}{R P} \cdot \frac{Q T}{P S}=\frac{C P}{C Q} \cdot \frac{C Q}{C P}=1 .
$$

Hence, $\operatorname{area}(R Q T)=\operatorname{area}(R S P)$.

Solution 2. Assume again $A C<B C$. Denote the circumcenter by $O$, and let $\gamma$ be the angle at $C$. Similarly to the first solution, from right triangles $C T Q$ and $C S P$ we obtain that $\angle O P Q=\angle O Q P=90^{\circ}-\frac{\gamma}{2}$. Then triangle $O P Q$ is isosceles, $O P=O Q$ and moreover $\angle P O Q=\gamma$.

As is well-known, point $R$ is the midpoint of arc $A B$ and $\angle R O A=\angle B O R=\gamma$.


Consider the rotation around point $O$ by angle $\gamma$. This transform moves $A$ to $R, R$ to $B$ and $Q$ to $P$; hence triangles $R Q A$ and $B P R$ are congruent and they have the same area.

Triangles $R Q T$ and $R Q A$ have $R Q$ as a common side, so the ratio between their areas is

$$
\frac{\operatorname{area}(R Q T)}{\operatorname{area}(R Q A)}=\frac{d(T, C R)}{d(A, C R)}=\frac{C T}{C A}=\frac{1}{2} .
$$

$(d(X, Y Z)$ denotes the distance between point $X$ and line $Y Z)$.
It can be obtained similarly that

$$
\frac{\operatorname{area}(R P S)}{\operatorname{area}(B P R)}=\frac{C S}{C B}=\frac{1}{2}
$$

Now the proof can be completed as

$$
\operatorname{area}(R Q T)=\frac{1}{2} \operatorname{area}(R Q A)=\frac{1}{2} \operatorname{area}(B P R)=\operatorname{area}(R P S) .
$$

G2. Given an isosceles triangle $A B C$ with $A B=A C$. The midpoint of side $B C$ is denoted by $M$. Let $X$ be a variable point on the shorter arc $M A$ of the circumcircle of triangle $A B M$. Let $T$ be the point in the angle domain $B M A$, for which $\angle T M X=90^{\circ}$ and $T X=B X$. Prove that $\angle M T B-\angle C T M$ does not depend on $X$.
(Canada)
Solution 1. Let $N$ be the midpoint of segment $B T$ (see Figure 1). Line $X N$ is the axis of symmetry in the isosceles triangle $B X T$, thus $\angle T N X=90^{\circ}$ and $\angle B X N=\angle N X T$. Moreover, in triangle $B C T$, line $M N$ is the midline parallel to $C T$; hence $\angle C T M=\angle N M T$.

Due to the right angles at points $M$ and $N$, these points lie on the circle with diameter $X T$. Therefore,

$$
\angle M T B=\angle M T N=\angle M X N \quad \text { and } \quad \angle C T M=\angle N M T=\angle N X T=\angle B X N
$$

Hence

$$
\angle M T B-\angle C T M=\angle M X N-\angle B X N=\angle M X B=\angle M A B
$$

which does not depend on $X$.


Figure 1


Figure 2

Solution 2. Let $S$ be the reflection of point $T$ over $M$ (see Figure 2). Then $X M$ is the perpendicular bisector of $T S$, hence $X B=X T=X S$, and $X$ is the circumcenter of triangle $B S T$. Moreover, $\angle B S M=\angle C T M$ since they are symmetrical about $M$. Then

$$
\angle M T B-\angle C T M=\angle S T B-\angle B S T=\frac{\angle S X B-\angle B X T}{2} .
$$

Observe that $\angle S X B=\angle S X T-\angle B X T=2 \angle M X T-\angle B X T$, so

$$
\angle M T B-\angle C T M=\frac{2 \angle M X T-2 \angle B X T}{2}=\angle M X B=\angle M A B,
$$

which is constant.

G3. The diagonals of a trapezoid $A B C D$ intersect at point $P$. Point $Q$ lies between the parallel lines $B C$ and $A D$ such that $\angle A Q D=\angle C Q B$, and line $C D$ separates points $P$ and $Q$. Prove that $\angle B Q P=\angle D A Q$.
(Ukraine)
Solution. Let $t=\frac{A D}{B C}$. Consider the homothety $h$ with center $P$ and scale $-t$. Triangles $P D A$ and $P B C$ are similar with ratio $t$, hence $h(B)=D$ and $h(C)=A$.


Let $Q^{\prime}=h(Q)$ (see Figure 1). Then points $Q, P$ and $Q^{\prime}$ are obviously collinear. Points $Q$ and $P$ lie on the same side of $A D$, as well as on the same side of $B C$; hence $Q^{\prime}$ and $P$ are also on the same side of $h(B C)=A D$, and therefore $Q$ and $Q^{\prime}$ are on the same side of $A D$. Moreover, points $Q$ and $C$ are on the same side of $B D$, while $Q^{\prime}$ and $A$ are on the opposite side (see Figure above).

By the homothety, $\angle A Q^{\prime} D=\angle C Q B=\angle A Q D$, hence quadrilateral $A Q^{\prime} Q D$ is cyclic. Then

$$
\angle D A Q=\angle D Q^{\prime} Q=\angle D Q^{\prime} P=\angle B Q P
$$

(the latter equality is valid by the homothety again).
Comment. The statement of the problem is a limit case of the following result.
In an arbitrary quadrilateral $A B C D$, let $P=A C \cap B D, I=A D \cap B C$, and let $Q$ be an arbitrary point which is not collinear with any two of points $A, B, C, D$. Then $\angle A Q D=\angle C Q B$ if and only if $\angle B Q P=\angle I Q A$ (angles are oriented; see Figure below to the left).

In the special case of the trapezoid, $I$ is an ideal point and $\angle D A Q=\angle I Q A=\angle B Q P$.


Let $a=Q A, b=Q B, c=Q C, d=Q D, i=Q I$ and $p=Q P$. Let line $Q A$ intersect lines $B C$ and $B D$ at points $U$ and $V$, respectively. On lines $B C$ and $B D$ we have

$$
(a b c i)=(U B C I) \quad \text { and } \quad(b a d p)=(a b p d)=(V B P D)
$$

Projecting from $A$, we get

$$
(a b c i)=(U B C I)=(V B P D)=(b a d p)
$$

Suppose that $\angle A Q D=\angle C Q B$. Let line $p^{\prime}$ be the reflection of line $i$ about the bisector of angle $A Q B$. Then by symmetry we have $\left(b a d p^{\prime}\right)=(a b c i)=(b a d p)$. Hence $p=p^{\prime}$, as desired.

The converse statement can be proved analogously.

G4. Consider five points $A, B, C, D, E$ such that $A B C D$ is a parallelogram and $B C E D$ is a cyclic quadrilateral. Let $\ell$ be a line passing through $A$, and let $\ell$ intersect segment $D C$ and line $B C$ at points $F$ and $G$, respectively. Suppose that $E F=E G=E C$. Prove that $\ell$ is the bisector of angle $D A B$.
(Luxembourg)
Solution. If $C F=C G$, then $\angle F G C=\angle G F C$, hence $\angle G A B=\angle G F C=\angle F G C=\angle F A D$, and $\ell$ is a bisector.

Assume that $C F<G C$. Let $E K$ and $E L$ be the altitudes in the isosceles triangles $E C F$ and $E G C$, respectively. Then in the right triangles $E K F$ and $E L C$ we have $E F=E C$ and

$$
K F=\frac{C F}{2}<\frac{G C}{2}=L C
$$

so

$$
K E=\sqrt{E F^{2}-K F^{2}}>\sqrt{E C^{2}-L C^{2}}=L E .
$$

Since quadrilateral $B C E D$ is cyclic, we have $\angle E D C=\angle E B C$, so the right triangles $B E L$ and $D E K$ are similar. Then $K E>L E$ implies $D K>B L$, and hence

$$
D F=D K-K F>B L-L C=B C=A D .
$$

But triangles $A D F$ and $G C F$ are similar, so we have $1>\frac{A D}{D F}=\frac{G C}{C F}$; this contradicts our assumption.

The case $C F>G C$ is completely similar. We consequently obtain the converse inequalities $K F>L C, K E<L E, D K<B L, D F<A D$, hence $1<\frac{A D}{D F}=\frac{G C}{C F}$; a contradiction.


G5. Let $A B C$ be a fixed triangle, and let $A_{1}, B_{1}, C_{1}$ be the midpoints of sides $B C, C A, A B$, respectively. Let $P$ be a variable point on the circumcircle. Let lines $P A_{1}, P B_{1}, P C_{1}$ meet the circumcircle again at $A^{\prime}, B^{\prime}, C^{\prime}$ respectively. Assume that the points $A, B, C, A^{\prime}, B^{\prime}, C^{\prime}$ are distinct, and lines $A A^{\prime}, B B^{\prime}, C C^{\prime}$ form a triangle. Prove that the area of this triangle does not depend on $P$.
(United Kingdom)
Solution 1. Let $A_{0}, B_{0}, C_{0}$ be the points of intersection of the lines $A A^{\prime}, B B^{\prime}$ and $C C^{\prime}$ (see Figure). We claim that area $\left(A_{0} B_{0} C_{0}\right)=\frac{1}{2}$ area $(A B C)$, hence it is constant.

Consider the inscribed hexagon $A B C C^{\prime} P A^{\prime}$. By Pascal's theorem, the points of intersection of its opposite sides (or of their extensions) are collinear. These points are $A B \cap C^{\prime} P=C_{1}$, $B C \cap P A^{\prime}=A_{1}, C C^{\prime} \cap A^{\prime} A=B_{0}$. So point $B_{0}$ lies on the midline $A_{1} C_{1}$ of triangle $A B C$. Analogously, points $A_{0}$ and $C_{0}$ lie on lines $B_{1} C_{1}$ and $A_{1} B_{1}$, respectively.

Lines $A C$ and $A_{1} C_{1}$ are parallel, so triangles $B_{0} C_{0} A_{1}$ and $A C_{0} B_{1}$ are similar; hence we have

$$
\frac{B_{0} C_{0}}{A C_{0}}=\frac{A_{1} C_{0}}{B_{1} C_{0}} .
$$

Analogously, from $B C \| B_{1} C_{1}$ we obtain

$$
\frac{A_{1} C_{0}}{B_{1} C_{0}}=\frac{B C_{0}}{A_{0} C_{0}}
$$

Combining these equalities, we get

$$
\frac{B_{0} C_{0}}{A C_{0}}=\frac{B C_{0}}{A_{0} C_{0}},
$$

or

$$
A_{0} C_{0} \cdot B_{0} C_{0}=A C_{0} \cdot B C_{0}
$$

Hence we have


$$
\operatorname{area}\left(A_{0} B_{0} C_{0}\right)=\frac{1}{2} A_{0} C_{0} \cdot B_{0} C_{0} \sin \angle A_{0} C_{0} B_{0}=\frac{1}{2} A C_{0} \cdot B C_{0} \sin \angle A C_{0} B=\operatorname{area}\left(A B C_{0}\right) .
$$

Since $C_{0}$ lies on the midline, we have $d\left(C_{0}, A B\right)=\frac{1}{2} d(C, A B)$ (we denote by $d(X, Y Z)$ the distance between point $X$ and line $Y Z$ ). Then we obtain

$$
\operatorname{area}\left(A_{0} B_{0} C_{0}\right)=\operatorname{area}\left(A B C_{0}\right)=\frac{1}{2} A B \cdot d\left(C_{0}, A B\right)=\frac{1}{4} A B \cdot d(C, A B)=\frac{1}{2} \operatorname{area}(A B C) .
$$

Solution 2. Again, we prove that area $\left(A_{0} B_{0} C_{0}\right)=\frac{1}{2}$ area $(A B C)$.
We can assume that $P$ lies on arc $A C$. Mark a point $L$ on side $A C$ such that $\angle C B L=$ $\angle P B A$; then $\angle L B A=\angle C B A-\angle C B L=\angle C B A-\angle P B A=\angle C B P$. Note also that $\angle B A L=\angle B A C=\angle B P C$ and $\angle L C B=\angle A P B$. Hence, triangles $B A L$ and $B P C$ are similar, and so are triangles $L C B$ and $A P B$.

Analogously, mark points $K$ and $M$ respectively on the extensions of sides $C B$ and $A B$ beyond point $B$, such that $\angle K A B=\angle C A P$ and $\angle B C M=\angle P C A$. For analogous reasons, $\angle K A C=\angle B A P$ and $\angle A C M=\angle P C B$. Hence $\triangle A B K \sim \triangle A P C \sim \triangle M B C, \triangle A C K \sim$ $\triangle A P B$, and $\triangle M A C \sim \triangle B P C$. From these similarities, we have $\angle C M B=\angle K A B=\angle C A P$, while we have seen that $\angle C A P=\angle C B P=\angle L B A$. Hence, $A K\|B L\| C M$.


Let line $C C^{\prime}$ intersect $B L$ at point $X$. Note that $\angle L C X=\angle A C C^{\prime}=\angle A P C^{\prime}=\angle A P C_{1}$, and $P C_{1}$ is a median in triangle $A P B$. Since triangles $A P B$ and $L C B$ are similar, $C X$ is a median in triangle $L C B$, and $X$ is a midpoint of $B L$. For the same reason, $A A^{\prime}$ passes through this midpoint, so $X=B_{0}$. Analogously, $A_{0}$ and $C_{0}$ are the midpoints of $A K$ and $C M$.

Now, from $A A_{0} \| C C_{0}$, we have

$$
\operatorname{area}\left(A_{0} B_{0} C_{0}\right)=\operatorname{area}\left(A C_{0} A_{0}\right)-\operatorname{area}\left(A B_{0} A_{0}\right)=\operatorname{area}\left(A C A_{0}\right)-\operatorname{area}\left(A B_{0} A_{0}\right)=\operatorname{area}\left(A C B_{0}\right) .
$$

Finally,

$$
\operatorname{area}\left(A_{0} B_{0} C_{0}\right)=\operatorname{area}\left(A C B_{0}\right)=\frac{1}{2} B_{0} L \cdot A C \sin A L B_{0}=\frac{1}{4} B L \cdot A C \sin A L B=\frac{1}{2} \operatorname{area}(A B C) .
$$

Comment 1. The equality area $\left(A_{0} B_{0} C_{0}\right)=\operatorname{area}\left(A C B_{0}\right)$ in Solution 2 does not need to be proved since the following fact is frequently known.

Suppose that the lines $K L$ and $M N$ are parallel, while the lines $K M$ and $L N$ intersect in a point $E$. Then $\operatorname{area}(K E N)=\operatorname{area}(M E L)$.
Comment 2. It follows immediately from both solutions that $A A_{0}\left\|B B_{0}\right\| C C_{0}$. These lines pass through an ideal point which is isogonally conjugate to $P$. It is known that they are parallel to the Simson line of point $Q$ which is opposite to $P$ on the circumcircle.
Comment 3. If $A=A^{\prime}$, then one can define the line $A A^{\prime}$ to be the tangent to the circumcircle at point $A$. Then the statement of the problem is also valid in this case.

G6. Determine the smallest positive real number $k$ with the following property.
Let $A B C D$ be a convex quadrilateral, and let points $A_{1}, B_{1}, C_{1}$ and $D_{1}$ lie on sides $A B, B C$, $C D$ and $D A$, respectively. Consider the areas of triangles $A A_{1} D_{1}, B B_{1} A_{1}, C C_{1} B_{1}$, and $D D_{1} C_{1}$; let $S$ be the sum of the two smallest ones, and let $S_{1}$ be the area of quadrilateral $A_{1} B_{1} C_{1} D_{1}$. Then we always have $k S_{1} \geq S$.

Answer. $k=1$.
Solution. Throughout the solution, triangles $A A_{1} D_{1}, B B_{1} A_{1}, C C_{1} B_{1}$, and $D D_{1} C_{1}$ will be referred to as border triangles. We will denote by $[\mathcal{R}]$ the area of a region $\mathcal{R}$.

First, we show that $k \geq 1$. Consider a triangle $A B C$ with unit area; let $A_{1}, B_{1}, K$ be the midpoints of its sides $A B, B C, A C$, respectively. Choose a point $D$ on the extension of $B K$, close to $K$. Take points $C_{1}$ and $D_{1}$ on sides $C D$ and $D A$ close to $D$ (see Figure 1). We have $\left[B B_{1} A_{1}\right]=\frac{1}{4}$. Moreover, as $C_{1}, D_{1}, D \rightarrow K$, we get $\left[A_{1} B_{1} C_{1} D_{1}\right] \rightarrow\left[A_{1} B_{1} K\right]=\frac{1}{4}$, $\left[A A_{1} D_{1}\right] \rightarrow\left[A A_{1} K\right]=\frac{1}{4},\left[C C_{1} B_{1}\right] \rightarrow\left[C K B_{1}\right]=\frac{1}{4}$ and $\left[D D_{1} C_{1}\right] \rightarrow 0$. Hence, the sum of the two smallest areas of border triangles tends to $\frac{1}{4}$, as well as $\left[A_{1} B_{1} C_{1} D_{1}\right]$; therefore, their ratio tends to 1 , and $k \geq 1$.

We are left to prove that $k=1$ satisfies the desired property.


Figure 1


Figure 2


Figure 3

Lemma. Let points $A_{1}, B_{1}, C_{1}$ lie respectively on sides $B C, C A, A B$ of a triangle $A B C$. Then $\left[A_{1} B_{1} C_{1}\right] \geq \min \left\{\left[A C_{1} B_{1}\right],\left[B A_{1} C_{1}\right],\left[C B_{1} A_{1}\right]\right\}$.
Proof. Let $A^{\prime}, B^{\prime}, C^{\prime}$ be the midpoints of sides $B C, C A$ and $A B$, respectively.
Suppose that two of points $A_{1}, B_{1}, C_{1}$ lie in one of triangles $A C^{\prime} B^{\prime}, B A^{\prime} C^{\prime}$ and $C B^{\prime} A^{\prime}$ (for convenience, let points $B_{1}$ and $C_{1}$ lie in triangle $A C^{\prime} B^{\prime}$; see Figure 2). Let segments $B_{1} C_{1}$ and $A A_{1}$ intersect at point $X$. Then $X$ also lies in triangle $A C^{\prime} B^{\prime}$. Hence $A_{1} X \geq A X$, and we have

$$
\frac{\left[A_{1} B_{1} C_{1}\right]}{\left[A C_{1} B_{1}\right]}=\frac{\frac{1}{2} A_{1} X \cdot B_{1} C_{1} \cdot \sin \angle A_{1} X C_{1}}{\frac{1}{2} A X \cdot B_{1} C_{1} \cdot \sin \angle A X B_{1}}=\frac{A_{1} X}{A X} \geq 1
$$

as required.
Otherwise, each one of triangles $A C^{\prime} B^{\prime}, B A^{\prime} C^{\prime}, C B^{\prime} A^{\prime}$ contains exactly one of points $A_{1}$, $B_{1}, C_{1}$, and we can assume that $B A_{1}<B A^{\prime}, C B_{1}<C B^{\prime}, A C_{1}<A C^{\prime}$ (see Figure 3). Then lines $B_{1} A_{1}$ and $A B$ intersect at a point $Y$ on the extension of $A B$ beyond point $B$, hence $\frac{\left[A_{1} B_{1} C_{1}\right]}{\left[A_{1} B_{1} C^{\prime}\right]}=\frac{C_{1} Y}{C^{\prime} Y}>1$; also, lines $A_{1} C^{\prime}$ and $C A$ intersect at a point $Z$ on the extension of $C A$ beyond point $A$, hence $\frac{\left[A_{1} B_{1} C^{\prime}\right]}{\left[A_{1} B^{\prime} C^{\prime}\right]}=\frac{B_{1} Z}{B^{\prime} Z}>1$. Finally, since $A_{1} A^{\prime} \| B^{\prime} C^{\prime}$, we have $\left[A_{1} B_{1} C_{1}\right]>\left[A_{1} B_{1} C^{\prime}\right]>\left[A_{1} B^{\prime} C^{\prime}\right]=\left[A^{\prime} B^{\prime} C^{\prime}\right]=\frac{1}{4}[A B C]$.

Now, from $\left[A_{1} B_{1} C_{1}\right]+\left[A C_{1} B_{1}\right]+\left[B A_{1} C_{1}\right]+\left[C B_{1} A_{1}\right]=[A B C]$ we obtain that one of the remaining triangles $A C_{1} B_{1}, B A_{1} C_{1}, C B_{1} A_{1}$ has an area less than $\frac{1}{4}[A B C]$, so it is less than $\left[A_{1} B_{1} C_{1}\right]$.

Now we return to the problem. We say that triangle $A_{1} B_{1} C_{1}$ is small if $\left[A_{1} B_{1} C_{1}\right]$ is less than each of $\left[B B_{1} A_{1}\right]$ and $\left[C C_{1} B_{1}\right]$; otherwise this triangle is big (the similar notion is introduced for triangles $B_{1} C_{1} D_{1}, C_{1} D_{1} A_{1}, D_{1} A_{1} B_{1}$ ). If both triangles $A_{1} B_{1} C_{1}$ and $C_{1} D_{1} A_{1}$ are big, then $\left[A_{1} B_{1} C_{1}\right]$ is not less than the area of some border triangle, and $\left[C_{1} D_{1} A_{1}\right]$ is not less than the area of another one; hence, $S_{1}=\left[A_{1} B_{1} C_{1}\right]+\left[C_{1} D_{1} A_{1}\right] \geq S$. The same is valid for the pair of $B_{1} C_{1} D_{1}$ and $D_{1} A_{1} B_{1}$. So it is sufficient to prove that in one of these pairs both triangles are big.

Suppose the contrary. Then there is a small triangle in each pair. Without loss of generality, assume that triangles $A_{1} B_{1} C_{1}$ and $D_{1} A_{1} B_{1}$ are small. We can assume also that $\left[A_{1} B_{1} C_{1}\right] \leq$ [ $D_{1} A_{1} B_{1}$ ]. Note that in this case ray $D_{1} C_{1}$ intersects line $B C$.

Consider two cases.


Figure 4


Figure 5

Case 1. Ray $C_{1} D_{1}$ intersects line $A B$ at some point $K$. Let ray $D_{1} C_{1}$ intersect line $B C$ at point $L$ (see Figure 4). Then we have $\left[A_{1} B_{1} C_{1}\right]<\left[C C_{1} B_{1}\right]<\left[L C_{1} B_{1}\right],\left[A_{1} B_{1} C_{1}\right]<\left[B B_{1} A_{1}\right]$ (both - since $\left[A_{1} B_{1} C_{1}\right]$ is small), and $\left[A_{1} B_{1} C_{1}\right] \leq\left[D_{1} A_{1} B_{1}\right]<\left[A A_{1} D_{1}\right]<\left[K A_{1} D_{1}\right]<\left[K A_{1} C_{1}\right]$ (since triangle $D_{1} A_{1} B_{1}$ is small). This contradicts the Lemma, applied for triangle $A_{1} B_{1} C_{1}$ inside $L K B$.

Case 2. Ray $C_{1} D_{1}$ does not intersect $A B$. Then choose a "sufficiently far" point $K$ on ray $B A$ such that $\left[K A_{1} C_{1}\right]>\left[A_{1} B_{1} C_{1}\right]$, and that ray $K C_{1}$ intersects line $B C$ at some point $L$ (see Figure 5). Since ray $C_{1} D_{1}$ does not intersect line $A B$, the points $A$ and $D_{1}$ are on different sides of $K L$; then $A$ and $D$ are also on different sides, and $C$ is on the same side as $A$ and $B$. Then analogously we have $\left[A_{1} B_{1} C_{1}\right]<\left[C C_{1} B_{1}\right]<\left[L C_{1} B_{1}\right]$ and $\left[A_{1} B_{1} C_{1}\right]<\left[B B_{1} A_{1}\right]$ since triangle $A_{1} B_{1} C_{1}$ is small. This (together with $\left[A_{1} B_{1} C_{1}\right]<\left[K A_{1} C_{1}\right]$ ) contradicts the Lemma again.

G7. Given an acute triangle $A B C$ with angles $\alpha, \beta$ and $\gamma$ at vertices $A, B$ and $C$, respectively, such that $\beta>\gamma$. Point $I$ is the incenter, and $R$ is the circumradius. Point $D$ is the foot of the altitude from vertex $A$. Point $K$ lies on line $A D$ such that $A K=2 R$, and $D$ separates $A$ and $K$. Finally, lines $D I$ and $K I$ meet sides $A C$ and $B C$ at $E$ and $F$, respectively.

Prove that if $I E=I F$ then $\beta \leq 3 \gamma$.

Solution 1. We first prove that

$$
\begin{equation*}
\angle K I D=\frac{\beta-\gamma}{2} \tag{1}
\end{equation*}
$$

even without the assumption that $I E=I F$. Then we will show that the statement of the problem is a consequence of this fact.

Denote the circumcenter by $O$. On the circumcircle, let $P$ be the point opposite to $A$, and let the angle bisector $A I$ intersect the circle again at $M$. Since $A K=A P=2 R$, triangle $A K P$ is isosceles. It is known that $\angle B A D=\angle C A O$, hence $\angle D A I=\angle B A I-\angle B A D=\angle C A I-$ $\angle C A O=\angle O A I$, and $A M$ is the bisector line in triangle $A K P$. Therefore, points $K$ and $P$ are symmetrical about $A M$, and $\angle A M K=\angle A M P=90^{\circ}$. Thus, $M$ is the midpoint of $K P$, and $A M$ is the perpendicular bisector of $K P$.


Denote the perpendicular feet of incenter $I$ on lines $B C, A C$, and $A D$ by $A_{1}, B_{1}$, and $T$, respectively. Quadrilateral $D A_{1} I T$ is a rectangle, hence $T D=I A_{1}=I B_{1}$.

Due to the right angles at $T$ and $B_{1}$, quadrilateral $A B_{1} I T$ is cyclic. Hence $\angle B_{1} T I=$ $\angle B_{1} A I=\angle C A M=\angle B A M=\angle B P M$ and $\angle I B_{1} T=\angle I A T=\angle M A K=\angle M A P=$ $\angle M B P$. Therefore, triangles $B_{1} T I$ and $B P M$ are similar and $\frac{I T}{I B_{1}}=\frac{M P}{M B}$.

It is well-known that $M B=M C=M I$. Then right triangles $I T D$ and $K M I$ are also
similar, because $\frac{I T}{T D}=\frac{I T}{I B_{1}}=\frac{M P}{M B}=\frac{K M}{M I}$. Hence, $\angle K I M=\angle I D T=\angle I D A$, and

$$
\angle K I D=\angle M I D-\angle K I M=(\angle I A D+\angle I D A)-\angle I D A=\angle I A D .
$$

Finally, from the right triangle $A D B$ we can compute

$$
\angle K I D=\angle I A D=\angle I A B-\angle D A B=\frac{\alpha}{2}-\left(90^{\circ}-\beta\right)=\frac{\alpha}{2}-\frac{\alpha+\beta+\gamma}{2}+\beta=\frac{\beta-\gamma}{2} .
$$

Now let us turn to the statement and suppose that $I E=I F$. Since $I A_{1}=I B_{1}$, the right triangles $I E B_{1}$ and $I F A_{1}$ are congruent and $\angle I E B_{1}=\angle I F A_{1}$. Since $\beta>\gamma, A_{1}$ lies in the interior of segment $C D$ and $F$ lies in the interior of $A_{1} D$. Hence, $\angle I F C$ is acute. Then two cases are possible depending on the order of points $A, C, B_{1}$ and $E$.


If point $E$ lies between $C$ and $B_{1}$ then $\angle I F C=\angle I E A$, hence quadrilateral $C E I F$ is cyclic and $\angle F C E=180^{\circ}-\angle E I F=\angle K I D$. By (1), in this case we obtain $\angle F C E=\gamma=\angle K I D=$ $\frac{\beta-\gamma}{2}$ and $\beta=3 \gamma$.

Otherwise, if point $E$ lies between $A$ and $B_{1}$, quadrilateral $C E I F$ is a deltoid such that $\angle I E C=\angle I F C<90^{\circ}$. Then we have $\angle F C E>180^{\circ}-\angle E I F=\angle K I D$. Therefore, $\angle F C E=\gamma>\angle K I D=\frac{\beta-\gamma}{2}$ and $\beta<3 \gamma$.
Comment 1. In the case when quadrilateral CEIF is a deltoid, one can prove the desired inequality without using (1). Actually, from $\angle I E C=\angle I F C<90^{\circ}$ it follows that $\angle A D I=90^{\circ}-\angle E D C<$ $\angle A E D-\angle E D C=\gamma$. Since the incircle lies inside triangle $A B C$, we have $A D>2 r$ (here $r$ is the inradius), which implies $D T<T A$ and $D I<A I$; hence $\frac{\beta-\gamma}{2}=\angle I A D<\angle A D I<\gamma$.
Solution 2. We give a different proof for (1). Then the solution can be finished in the same way as above.

Define points $M$ and $P$ again; it can be proved in the same way that $A M$ is the perpendicular bisector of $K P$. Let $J$ be the center of the excircle touching side $B C$. It is well-known that points $B, C, I, J$ lie on a circle with center $M$; denote this circle by $\omega_{1}$.

Let $B^{\prime}$ be the reflection of point $B$ about the angle bisector $A M$. By the symmetry, $B^{\prime}$ is the second intersection point of circle $\omega_{1}$ and line $A C$. Triangles $P B A$ and $K B^{\prime} A$ are symmetrical
with respect to line $A M$, therefore $\angle K B^{\prime} A=\angle P B A=90^{\circ}$. By the right angles at $D$ and $B^{\prime}$, points $K, D, B^{\prime}, C$ are concyclic and

$$
A D \cdot A K=A B^{\prime} \cdot A C
$$

From the cyclic quadrilateral $I J C B^{\prime}$ we obtain $A B^{\prime} \cdot A C=A I \cdot A J$ as well, therefore

$$
A D \cdot A K=A B^{\prime} \cdot A C=A I \cdot A J
$$

and points $I, J, K, D$ are also concyclic. Denote circle $I D K J$ by $\omega_{2}$.


Let $N$ be the point on circle $\omega_{2}$ which is opposite to $K$. Since $\angle N D K=90^{\circ}=\angle C D K$, point $N$ lies on line $B C$. Point $M$, being the center of circle $\omega_{1}$, is the midpoint of segment $I J$, and $K M$ is perpendicular to $I J$. Therefore, line $K M$ is the perpendicular bisector of $I J$ and hence it passes through $N$.

From the cyclic quadrilateral $I D K N$ we obtain

$$
\angle K I D=\angle K N D=90^{\circ}-\angle D K N=90^{\circ}-\angle A K M=\angle M A K=\frac{\beta-\gamma}{2} .
$$

Comment 2. The main difficulty in the solution is finding (1). If someone can guess this fact, he or she can compute it in a relatively short way.

One possible way is finding and applying the relation $A I^{2}=2 R\left(h_{a}-2 r\right)$, where $h_{a}=A D$ is the length of the altitude. Using this fact, one can see that triangles $A K I$ and $A I D^{\prime}$ are similar (here $D^{\prime}$ is the point symmetrical to $D$ about $T$ ). Hence, $\angle M I K=\angle D D^{\prime} I=\angle I D D^{\prime}$. The proof can be finished as in Solution 1.

G8. Point $P$ lies on side $A B$ of a convex quadrilateral $A B C D$. Let $\omega$ be the incircle of triangle $C P D$, and let $I$ be its incenter. Suppose that $\omega$ is tangent to the incircles of triangles $A P D$ and $B P C$ at points $K$ and $L$, respectively. Let lines $A C$ and $B D$ meet at $E$, and let lines $A K$ and $B L$ meet at $F$. Prove that points $E, I$, and $F$ are collinear.
(Poland)
Solution. Let $\Omega$ be the circle tangent to segment $A B$ and to rays $A D$ and $B C$; let $J$ be its center. We prove that points $E$ and $F$ lie on line $I J$.


Denote the incircles of triangles $A D P$ and $B C P$ by $\omega_{A}$ and $\omega_{B}$. Let $h_{1}$ be the homothety with a negative scale taking $\omega$ to $\Omega$. Consider this homothety as the composition of two homotheties: one taking $\omega$ to $\omega_{A}$ (with a negative scale and center $K$ ), and another one taking $\omega_{A}$ to $\Omega$ (with a positive scale and center $A$ ). It is known that in such a case the three centers of homothety are collinear (this theorem is also referred to as the theorem on the three similitude centers). Hence, the center of $h_{1}$ lies on line $A K$. Analogously, it also lies on $B L$, so this center is $F$. Hence, $F$ lies on the line of centers of $\omega$ and $\Omega$, i. e. on $I J$ (if $I=J$, then $F=I$ as well, and the claim is obvious).

Consider quadrilateral $A P C D$ and mark the equal segments of tangents to $\omega$ and $\omega_{A}$ (see the figure below to the left). Since circles $\omega$ and $\omega_{A}$ have a common point of tangency with $P D$, one can easily see that $A D+P C=A P+C D$. So, quadrilateral $A P C D$ is circumscribed; analogously, circumscribed is also quadrilateral $B C D P$. Let $\Omega_{A}$ and $\Omega_{B}$ respectively be their incircles.


Consider the homothety $h_{2}$ with a positive scale taking $\omega$ to $\Omega$. Consider $h_{2}$ as the composition of two homotheties: taking $\omega$ to $\Omega_{A}$ (with a positive scale and center $C$ ), and taking $\Omega_{A}$ to $\Omega$ (with a positive scale and center $A$ ), respectively. So the center of $h_{2}$ lies on line $A C$. By analogous reasons, it lies also on $B D$, hence this center is $E$. Thus, $E$ also lies on the line of centers $I J$, and the claim is proved.
Comment. In both main steps of the solution, there can be several different reasonings for the same claims. For instance, one can mostly use Desargues' theorem instead of the three homotheties theorem. Namely, if $I_{A}$ and $I_{B}$ are the centers of $\omega_{A}$ and $\omega_{B}$, then lines $I_{A} I_{B}, K L$ and $A B$ are concurrent (by the theorem on three similitude centers applied to $\omega, \omega_{A}$ and $\omega_{B}$ ). Then Desargues' theorem, applied to triangles $A I_{A} K$ and $B I_{B} L$, yields that the points $J=A I_{A} \cap B I_{B}, I=I_{A} K \cap I_{B} L$ and $F=A K \cap B L$ are collinear.

For the second step, let $J_{A}$ and $J_{B}$ be the centers of $\Omega_{A}$ and $\Omega_{B}$. Then lines $J_{A} J_{B}, A B$ and $C D$ are concurrent, since they appear to be the two common tangents and the line of centers of $\Omega_{A}$ and $\Omega_{B}$. Applying Desargues' theorem to triangles $A J_{A} C$ and $B J_{B} D$, we obtain that the points $J=A J_{A} \cap B J_{B}$, $I=C J_{A} \cap D J_{B}$ and $E=A C \cap B D$ are collinear.

## Number Theory

N1. Find all pairs $(k, n)$ of positive integers for which $7^{k}-3^{n}$ divides $k^{4}+n^{2}$.
(Austria)
Answer. (2, 4).
Solution. Suppose that a pair $(k, n)$ satisfies the condition of the problem. Since $7^{k}-3^{n}$ is even, $k^{4}+n^{2}$ is also even, hence $k$ and $n$ have the same parity. If $k$ and $n$ are odd, then $k^{4}+n^{2} \equiv 1+1=2(\bmod 4)$, while $7^{k}-3^{n} \equiv 7-3 \equiv 0(\bmod 4)$, so $k^{4}+n^{2}$ cannot be divisible by $7^{k}-3^{n}$. Hence, both $k$ and $n$ must be even.

Write $k=2 a, n=2 b$. Then $7^{k}-3^{n}=7^{2 a}-3^{2 b}=\frac{7^{a}-3^{b}}{2} \cdot 2\left(7^{a}+3^{b}\right)$, and both factors are integers. So $2\left(7^{a}+3^{b}\right) \mid 7^{k}-3^{n}$ and $7^{k}-3^{n} \mid k^{4}+n^{2}=2\left(8 a^{4}+2 b^{2}\right)$, hence

$$
\begin{equation*}
7^{a}+3^{b} \leq 8 a^{4}+2 b^{2} \tag{1}
\end{equation*}
$$

We prove by induction that $8 a^{4}<7^{a}$ for $a \geq 4,2 b^{2}<3^{b}$ for $b \geq 1$ and $2 b^{2}+9 \leq 3^{b}$ for $b \geq 3$. In the initial cases $a=4, b=1, b=2$ and $b=3$ we have $8 \cdot 4^{4}=2048<7^{4}=2401,2<3$, $2 \cdot 2^{2}=8<3^{2}=9$ and $2 \cdot 3^{2}+9=3^{3}=27$, respectively.

If $8 a^{4}<7^{a}(a \geq 4)$ and $2 b^{2}+9 \leq 3^{b}(b \geq 3)$, then

$$
\begin{aligned}
8(a+1)^{4} & =8 a^{4}\left(\frac{a+1}{a}\right)^{4}<7^{a}\left(\frac{5}{4}\right)^{4}=7^{a} \frac{625}{256}<7^{a+1} \quad \text { and } \\
2(b+1)^{2}+9 & <\left(2 b^{2}+9\right)\left(\frac{b+1}{b}\right)^{2} \leq 3^{b}\left(\frac{4}{3}\right)^{2}=3^{b} \frac{16}{9}<3^{b+1},
\end{aligned}
$$

as desired.
For $a \geq 4$ we obtain $7^{a}+3^{b}>8 a^{4}+2 b^{2}$ and inequality (1) cannot hold. Hence $a \leq 3$, and three cases are possible.

Case 1: $a=1$. Then $k=2$ and $8+2 b^{2} \geq 7+3^{b}$, thus $2 b^{2}+1 \geq 3^{b}$. This is possible only if $b \leq 2$. If $b=1$ then $n=2$ and $\frac{k^{4}+n^{2}}{7^{k}-3^{n}}=\frac{2^{4}+2^{2}}{7^{2}-3^{2}}=\frac{1}{2}$, which is not an integer. If $b=2$ then $n=4$ and $\frac{k^{4}+n^{2}}{7^{k}-3^{n}}=\frac{2^{4}+4^{2}}{7^{2}-3^{4}}=-1$, so $(k, n)=(2,4)$ is a solution.

Case 2: $a=2$. Then $k=4$ and $k^{4}+n^{2}=256+4 b^{2} \geq\left|7^{4}-3^{n}\right|=\left|49-3^{b}\right| \cdot\left(49+3^{b}\right)$. The smallest value of the first factor is 22 , attained at $b=3$, so $128+2 b^{2} \geq 11\left(49+3^{b}\right)$, which is impossible since $3^{b}>2 b^{2}$.

Case 3: $a=3$. Then $k=6$ and $k^{4}+n^{2}=1296+4 b^{2} \geq\left|7^{6}-3^{n}\right|=\left|343-3^{b}\right| \cdot\left(343+3^{b}\right)$. Analogously, $\left|343-3^{b}\right| \geq 100$ and we have $324+b^{2} \geq 25\left(343+3^{b}\right)$, which is impossible again.

We find that there exists a unique solution $(k, n)=(2,4)$.

N2. Let $b, n>1$ be integers. Suppose that for each $k>1$ there exists an integer $a_{k}$ such that $b-a_{k}^{n}$ is divisible by $k$. Prove that $b=A^{n}$ for some integer $A$.
(Canada)
Solution. Let the prime factorization of $b$ be $b=p_{1}^{\alpha_{1}} \ldots p_{s}^{\alpha_{s}}$, where $p_{1}, \ldots, p_{s}$ are distinct primes. Our goal is to show that all exponents $\alpha_{i}$ are divisible by $n$, then we can set $A=p_{1}^{\alpha_{1} / n} \ldots p_{s}^{\alpha_{s} / n}$.

Apply the condition for $k=b^{2}$. The number $b-a_{k}^{n}$ is divisible by $b^{2}$ and hence, for each $1 \leq i \leq s$, it is divisible by $p_{i}^{2 \alpha_{i}}>p_{i}^{\alpha_{i}}$ as well. Therefore

$$
a_{k}^{n} \equiv b \equiv 0 \quad\left(\bmod p_{i}^{\alpha_{i}}\right)
$$

and

$$
a_{k}^{n} \equiv b \not \equiv 0 \quad\left(\bmod p_{i}^{\alpha_{i}+1}\right),
$$

which implies that the largest power of $p_{i}$ dividing $a_{k}^{n}$ is $p_{i}^{\alpha_{i}}$. Since $a_{k}^{n}$ is a complete $n$th power, this implies that $\alpha_{i}$ is divisible by $n$.
Comment. If $n=8$ and $b=16$, then for each prime $p$ there exists an integer $a_{p}$ such that $b-a_{p}^{n}$ is divisible by $p$. Actually, the congruency $x^{8}-16 \equiv 0(\bmod p)$ expands as

$$
\left(x^{2}-2\right)\left(x^{2}+2\right)\left(x^{2}-2 x+2\right)\left(x^{2}+2 x+2\right) \equiv 0 \quad(\bmod p) .
$$

Hence, if -1 is a quadratic residue modulo $p$, then congruency $x^{2}+2 x+2=(x+1)^{2}+1 \equiv 0$ has a solution. Otherwise, one of congruencies $x^{2} \equiv 2$ and $x^{2} \equiv-2$ has a solution.

Thus, the solution cannot work using only prime values of $k$.

N3. Let $X$ be a set of 10000 integers, none of them is divisible by 47. Prove that there exists a 2007-element subset $Y$ of $X$ such that $a-b+c-d+e$ is not divisible by 47 for any $a, b, c, d, e \in Y$.
(Netherlands)
Solution. Call a set $M$ of integers good if $47 \nmid a-b+c-d+e$ for any $a, b, c, d, e \in M$.
Consider the set $J=\{-9,-7,-5,-3,-1,1,3,5,7,9\}$. We claim that $J$ is good. Actually, for any $a, b, c, d, e \in J$ the number $a-b+c-d+e$ is odd and

$$
-45=(-9)-9+(-9)-9+(-9) \leq a-b+c-d+e \leq 9-(-9)+9-(-9)+9=45
$$

But there is no odd number divisible by 47 between -45 and 45 .
For any $k=1, \ldots, 46$ consider the set

$$
A_{k}=\{x \in X \mid \exists j \in J: \quad k x \equiv j(\bmod 47)\} .
$$

If $A_{k}$ is not good, then $47 \mid a-b+c-d+e$ for some $a, b, c, d, e \in A_{k}$, hence $47 \mid k a-k b+$ $k c-k d+k e$. But set $J$ contains numbers with the same residues modulo 47, so $J$ also is not good. This is a contradiction; therefore each $A_{k}$ is a good subset of $X$.

Then it suffices to prove that there exists a number $k$ such that $\left|A_{k}\right| \geq 2007$. Note that each $x \in X$ is contained in exactly 10 sets $A_{k}$. Then

$$
\sum_{k=1}^{46}\left|A_{k}\right|=10|X|=100000
$$

hence for some value of $k$ we have

$$
\left|A_{k}\right| \geq \frac{100000}{46}>2173>2007
$$

This completes the proof.
Comment. For the solution, it is essential to find a good set consisting of 10 different residues. Actually, consider a set $X$ containing almost uniform distribution of the nonzero residues (i.e. each residue occurs 217 or 218 times). Let $Y \subset X$ be a good subset containing 2007 elements. Then the set $K$ of all residues appearing in $Y$ contains not less than 10 residues, and obviously this set is good.

On the other hand, there is no good set $K$ consisting of 11 different residues. The CauchyDavenport theorem claims that for any sets $A, B$ of residues modulo a prime $p$,

$$
|A+B| \geq \min \{p,|A|+|B|-1\} .
$$

Hence, if $|K| \geq 11$, then $|K+K| \geq 21,|K+K+K| \geq 31>47-|K+K|$, hence $\mid K+K+K+$ $(-K)+(-K) \mid=47$, and $0 \equiv a+c+e-b-d(\bmod 47)$ for some $a, b, c, d, e \in K$.

From the same reasoning, one can see that a good set $K$ containing 10 residues should satisfy equalities $|K+K|=19=2|K|-1$ and $|K+K+K|=28=|K+K|+|K|-1$. It can be proved that in this case set $K$ consists of 10 residues forming an arithmetic progression. As an easy consequence, one obtains that set $K$ has the form $a J$ for some nonzero residue $a$.
$\mathbf{N} 4$. For every integer $k \geq 2$, prove that $2^{3 k}$ divides the number

$$
\begin{equation*}
\binom{2^{k+1}}{2^{k}}-\binom{2^{k}}{2^{k-1}} \tag{1}
\end{equation*}
$$

but $2^{3 k+1}$ does not.
(Poland)
Solution. We use the notation $(2 n-1)!!=1 \cdot 3 \cdots(2 n-1)$ and $(2 n)!!=2 \cdot 4 \cdots(2 n)=2^{n} n!$ for any positive integer $n$. Observe that $(2 n)!=(2 n)!!(2 n-1)!!=2^{n} n!(2 n-1)!!$.

For any positive integer $n$ we have

$$
\begin{aligned}
& \binom{4 n}{2 n}=\frac{(4 n)!}{(2 n)!^{2}}=\frac{2^{2 n}(2 n)!(4 n-1)!!}{(2 n)!^{2}}=\frac{2^{2 n}}{(2 n)!}(4 n-1)!! \\
& \binom{2 n}{n}=\frac{1}{(2 n)!}\left(\frac{(2 n)!}{n!}\right)^{2}=\frac{1}{(2 n)!}\left(2^{n}(2 n-1)!!\right)^{2}=\frac{2^{2 n}}{(2 n)!}(2 n-1)!^{2}
\end{aligned}
$$

Then expression (1) can be rewritten as follows:

$$
\begin{align*}
\binom{2^{k+1}}{2^{k}} & -\binom{2^{k}}{2^{k-1}}=\frac{2^{2^{k}}}{\left(2^{k}\right)!}\left(2^{k+1}-1\right)!!-\frac{2^{2^{k}}}{\left(2^{k}\right)!}\left(2^{k}-1\right)!!^{2} \\
& =\frac{2^{2^{k}}\left(2^{k}-1\right)!!}{\left(2^{k}\right)!} \cdot\left(\left(2^{k}+1\right)\left(2^{k}+3\right) \ldots\left(2^{k}+2^{k}-1\right)-\left(2^{k}-1\right)\left(2^{k}-3\right) \ldots\left(2^{k}-2^{k}+1\right)\right) \tag{2}
\end{align*}
$$

We compute the exponent of 2 in the prime decomposition of each factor (the first one is a rational number but not necessarily an integer; it is not important).

First, we show by induction on $n$ that the exponent of 2 in $\left(2^{n}\right)$ ! is $2^{n}-1$. The base case $n=1$ is trivial. Suppose that $\left(2^{n}\right)!=2^{2^{n}-1}(2 d+1)$ for some integer $d$. Then we have

$$
\left(2^{n+1}\right)!=2^{2^{n}}\left(2^{n}\right)!\left(2^{n+1}-1\right)!!=2^{2^{n}} 2^{2^{n}-1} \cdot(2 d+1)\left(2^{n+1}-1\right)!!=2^{2^{n+1}-1} \cdot(2 q+1)
$$

for some integer $q$. This finishes the induction step.
Hence, the exponent of 2 in the first factor in $(2)$ is $2^{k}-\left(2^{k}-1\right)=1$.
The second factor in (2) can be considered as the value of the polynomial

$$
\begin{equation*}
P(x)=(x+1)(x+3) \ldots\left(x+2^{k}-1\right)-(x-1)(x-3) \ldots\left(x-2^{k}+1\right) . \tag{3}
\end{equation*}
$$

at $x=2^{k}$. Now we collect some information about $P(x)$.
Observe that $P(-x)=-P(x)$, since $k \geq 2$. So $P(x)$ is an odd function, and it has nonzero coefficients only at odd powers of $x$. Hence $P(x)=x^{3} Q(x)+c x$, where $Q(x)$ is a polynomial with integer coefficients.

Compute the exponent of 2 in $c$. We have

$$
\begin{aligned}
c & =2\left(2^{k}-1\right)!!\sum_{i=1}^{2^{k-1}} \frac{1}{2 i-1}=\left(2^{k}-1\right)!!\sum_{i=1}^{2^{k-1}}\left(\frac{1}{2 i-1}+\frac{1}{2^{k}-2 i+1}\right) \\
& =\left(2^{k}-1\right)!!\sum_{i=1}^{2^{k-1}} \frac{2^{k}}{(2 i-1)\left(2^{k}-2 i+1\right)}=2^{k} \sum_{i=1}^{2^{k-1}} \frac{\left(2^{k}-1\right)!!}{(2 i-1)\left(2^{k}-2 i+1\right)}=2^{k} S
\end{aligned}
$$

For any integer $i=1, \ldots, 2^{k-1}$, denote by $a_{2 i-1}$ the residue inverse to $2 i-1$ modulo $2^{k}$. Clearly, when $2 i-1$ runs through all odd residues, so does $a_{2 i-1}$, hence

$$
\begin{aligned}
S=\sum_{i=1}^{2^{k-1}} \frac{\left(2^{k}-1\right)!!}{(2 i-1)\left(2^{k}-2 i+1\right)} \equiv-\sum_{i=1}^{2^{k-1}} \frac{\left(2^{k}-1\right)!!}{(2 i-1)^{2}} \equiv-\sum_{i=1}^{2^{k-1}}\left(2^{k}-1\right)!!a_{2 i-1}^{2} \\
=-\left(2^{k}-1\right)!!\sum_{i=1}^{2^{k-1}}(2 i-1)^{2}=-\left(2^{k}-1\right)!!\frac{2^{k-1}\left(2^{2 k}-1\right)}{3} \quad\left(\bmod 2^{k}\right)
\end{aligned}
$$

Therefore, the exponent of 2 in $S$ is $k-1$, so $c=2^{k} S=2^{2 k-1}(2 t+1)$ for some integer $t$.
Finally we obtain that

$$
P\left(2^{k}\right)=2^{3 k} Q\left(2^{k}\right)+2^{k} c=2^{3 k} Q\left(2^{k}\right)+2^{3 k-1}(2 t+1),
$$

which is divisible exactly by $2^{3 k-1}$. Thus, the exponent of 2 in $(2)$ is $1+(3 k-1)=3 k$.
Comment. The fact that (1) is divisible by $2^{2 k}$ is known; but it does not help in solving this problem.

N5. Find all surjective functions $f: \mathbb{N} \rightarrow \mathbb{N}$ such that for every $m, n \in \mathbb{N}$ and every prime $p$, the number $f(m+n)$ is divisible by $p$ if and only if $f(m)+f(n)$ is divisible by $p$.
( $\mathbb{N}$ is the set of all positive integers.)
(Iran)
Answer. $f(n)=n$.
Solution. Suppose that function $f: \mathbb{N} \rightarrow \mathbb{N}$ satisfies the problem conditions.
Lemma. For any prime $p$ and any $x, y \in \mathbb{N}$, we have $x \equiv y(\bmod p)$ if and only if $f(x) \equiv f(y)$ $(\bmod p)$. Moreover, $p \mid f(x)$ if and only if $p \mid x$.
Proof. Consider an arbitrary prime $p$. Since $f$ is surjective, there exists some $x \in \mathbb{N}$ such that $p \mid f(x)$. Let

$$
d=\min \{x \in \mathbb{N}: p \mid f(x)\} .
$$

By induction on $k$, we obtain that $p \mid f(k d)$ for all $k \in \mathbb{N}$. The base is true since $p \mid f(d)$. Moreover, if $p \mid f(k d)$ and $p \mid f(d)$ then, by the problem condition, $p \mid f(k d+d)=f((k+1) d)$ as required.

Suppose that there exists an $x \in \mathbb{N}$ such that $d \nmid x$ but $p \mid f(x)$. Let

$$
y=\min \{x \in \mathbb{N}: d \nmid x, p \mid f(x)\} .
$$

By the choice of $d$, we have $y>d$, and $y-d$ is a positive integer not divisible by $d$. Then $p \nmid f(y-d)$, while $p \mid f(d)$ and $p \mid f(d+(y-d))=f(y)$. This contradicts the problem condition. Hence, there is no such $x$, and

$$
\begin{equation*}
p|f(x) \Longleftrightarrow d| x \tag{1}
\end{equation*}
$$

Take arbitrary $x, y \in \mathbb{N}$ such that $x \equiv y(\bmod d)$. We have $p \mid f(x+(2 x d-x))=f(2 x d)$; moreover, since $d \mid 2 x d+(y-x)=y+(2 x d-x)$, we get $p \mid f(y+(2 x d-x))$. Then by the problem condition $p|f(x)+f(2 x d-x), p| f(y)+f(2 x d-x)$, and hence $f(x) \equiv-f(2 x d-x) \equiv f(y)$ $(\bmod p)$.

On the other hand, assume that $f(x) \equiv f(y)(\bmod p)$. Again we have $p \mid f(x)+f(2 x d-x)$ which by our assumption implies that $p \mid f(x)+f(2 x d-x)+(f(y)-f(x))=f(y)+f(2 x d-x)$. Hence by the problem condition $p \mid f(y+(2 x d-x))$. Using (1) we get $0 \equiv y+(2 x d-x) \equiv y-x$ $(\bmod d)$.

Thus, we have proved that

$$
\begin{equation*}
x \equiv y \quad(\bmod d) \Longleftrightarrow f(x) \equiv f(y) \quad(\bmod p) \tag{2}
\end{equation*}
$$

We are left to show that $p=d$ : in this case (1) and (2) provide the desired statements.
The numbers $1,2, \ldots, d$ have distinct residues modulo $d$. By (2), numbers $f(1), f(2), \ldots$, $f(d)$ have distinct residues modulo $p$; hence there are at least $d$ distinct residues, and $p \geq d$. On the other hand, by the surjectivity of $f$, there exist $x_{1}, \ldots, x_{p} \in \mathbb{N}$ such that $f\left(x_{i}\right)=i$ for any $i=1,2, \ldots, p$. By (2), all these $x_{i}$ 's have distinct residues modulo $d$. For the same reasons, $d \geq p$. Hence, $d=p$.

Now we prove that $f(n)=n$ by induction on $n$. If $n=1$ then, by the Lemma, $p \nmid f(1)$ for any prime $p$, so $f(1)=1$, and the base is established. Suppose that $n>1$ and denote $k=f(n)$. Note that there exists a prime $q \mid n$, so by the Lemma $q \mid k$ and $k>1$.

If $k>n$ then $k-n+1>1$, and there exists a prime $p \mid k-n+1$; we have $k \equiv n-1$ $(\bmod p)$. By the induction hypothesis we have $f(n-1)=n-1 \equiv k=f(n)(\bmod p)$. Now, by the Lemma we obtain $n-1 \equiv n(\bmod p)$ which cannot be true.

Analogously, if $k<n$, then $f(k-1)=k-1$ by induction hypothesis. Moreover, $n-k+1>1$, so there exists a prime $p \mid n-k+1$ and $n \equiv k-1(\bmod p)$. By the Lemma again, $k=f(n) \equiv$ $f(k-1)=k-1(\bmod p)$, which is also false. The only remaining case is $k=n$, so $f(n)=n$.

Finally, the function $f(n)=n$ obviously satisfies the condition.

N6. Let $k$ be a positive integer. Prove that the number $\left(4 k^{2}-1\right)^{2}$ has a positive divisor of the form $8 k n-1$ if and only if $k$ is even.
(United Kingdom)
Solution. The statement follows from the following fact.
Lemma. For arbitrary positive integers $x$ and $y$, the number $4 x y-1$ divides $\left(4 x^{2}-1\right)^{2}$ if and only if $x=y$.
Proof. If $x=y$ then $4 x y-1=4 x^{2}-1$ obviously divides $\left(4 x^{2}-1\right)^{2}$ so it is sufficient to consider the opposite direction.

Call a pair $(x, y)$ of positive integers bad if $4 x y-1$ divides $\left(4 x^{2}-1\right)^{2}$ but $x \neq y$. In order to prove that bad pairs do not exist, we present two properties of them which provide an infinite descent.
Property (i). If $(x, y)$ is a bad pair and $x<y$ then there exists a positive integer $z<x$ such that $(x, z)$ is also bad.
Let $r=\frac{\left(4 x^{2}-1\right)^{2}}{4 x y-1}$. Then

$$
r=-r \cdot(-1) \equiv-r(4 x y-1)=-\left(4 x^{2}-1\right)^{2} \equiv-1 \quad(\bmod 4 x)
$$

and $r=4 x z-1$ with some positive integer $z$. From $x<y$ we obtain that

$$
4 x z-1=\frac{\left(4 x^{2}-1\right)^{2}}{4 x y-1}<4 x^{2}-1
$$

and therefore $z<x$. By the construction, the number $4 x z-1$ is a divisor of $\left(4 x^{2}-1\right)^{2}$ so $(x, z)$ is a bad pair.
Property (ii). If $(x, y)$ is a bad pair then $(y, x)$ is also bad.
Since $1=1^{2} \equiv(4 x y)^{2}(\bmod 4 x y-1)$, we have

$$
\left(4 y^{2}-1\right)^{2} \equiv\left(4 y^{2}-(4 x y)^{2}\right)^{2}=16 y^{4}\left(4 x^{2}-1\right)^{2} \equiv 0 \quad(\bmod 4 x y-1)
$$

Hence, the number $4 x y-1$ divides $\left(4 y^{2}-1\right)^{2}$ as well.
Now suppose that there exists at least one bad pair. Take a bad pair $(x, y)$ such that $2 x+y$ attains its smallest possible value. If $x<y$ then property (i) provides a bad pair $(x, z)$ with $z<y$ and thus $2 x+z<2 x+y$. Otherwise, if $y<x$, property (ii) yields that pair $(y, x)$ is also bad while $2 y+x<2 x+y$. Both cases contradict the assumption that $2 x+y$ is minimal; the Lemma is proved.

To prove the problem statement, apply the Lemma for $x=k$ and $y=2 n$; the number $8 k n-1$ divides $\left(4 k^{2}-1\right)^{2}$ if and only if $k=2 n$. Hence, there is no such $n$ if $k$ is odd and $n=k / 2$ is the only solution if $k$ is even.
Comment. The constant 4 in the Lemma can be replaced with an arbitrary integer greater than 1 : if $a>1$ and $a x y-1$ divides $\left(a x^{2}-1\right)^{2}$ then $x=y$.

N7. For a prime $p$ and a positive integer $n$, denote by $\nu_{p}(n)$ the exponent of $p$ in the prime factorization of $n$ !. Given a positive integer $d$ and a finite set $\left\{p_{1}, \ldots, p_{k}\right\}$ of primes. Show that there are infinitely many positive integers $n$ such that $d \mid \nu_{p_{i}}(n)$ for all $1 \leq i \leq k$.
(India)
Solution 1. For arbitrary prime $p$ and positive integer $n$, denote by $\operatorname{ord}_{p}(n)$ the exponent of $p$ in $n$. Thus,

$$
\nu_{p}(n)=\operatorname{ord}_{p}(n!)=\sum_{i=1}^{n} \operatorname{ord}_{p}(i)
$$

Lemma. Let $p$ be a prime number, $q$ be a positive integer, $k$ and $r$ be positive integers such that $p^{k}>r$. Then $\nu_{p}\left(q p^{k}+r\right)=\nu_{p}\left(q p^{k}\right)+\nu_{p}(r)$.
Proof. We claim that $\operatorname{ord}_{p}\left(q p^{k}+i\right)=\operatorname{ord}_{p}(i)$ for all $0<i<p^{k}$. Actually, if $d=\operatorname{ord}_{p}(i)$ then $d<k$, so $q p^{k}+i$ is divisible by $p^{d}$, but only the first term is divisible by $p^{d+1}$; hence the sum is not.

Using this claim, we obtain

$$
\nu_{p}\left(q p^{k}+r\right)=\sum_{i=1}^{q p^{k}} \operatorname{ord}_{p}(i)+\sum_{i=q p^{k}+1}^{q p^{k}+r} \operatorname{ord}_{p}(i)=\sum_{i=1}^{q p^{k}} \operatorname{ord}_{p}(i)+\sum_{i=1}^{r} \operatorname{ord}_{p}(i)=\nu_{p}\left(q p^{k}\right)+\nu_{p}(r)
$$

For any integer $a$, denote by $\bar{a}$ its residue modulo $d$. The addition of residues will also be performed modulo $d$, i. e. $\bar{a}+\bar{b}=\overline{a+b}$. For any positive integer $n$, let $f(n)=\left(f_{1}(n), \ldots, f_{k}(n)\right)$, where $f_{i}(n)=\overline{\nu_{p_{i}}(n)}$.

Define the sequence $n_{1}=1, n_{\ell+1}=\left(p_{1} p_{2} \ldots p_{k}\right)^{n_{\ell}}$. We claim that

$$
f\left(n_{\ell_{1}}+n_{\ell_{2}}+\ldots+n_{\ell_{m}}\right)=f\left(n_{\ell_{1}}\right)+f\left(n_{\ell_{2}}\right)+\ldots+f\left(n_{\ell_{m}}\right)
$$

for any $\ell_{1}<\ell_{2}<\ldots<\ell_{m}$. (The addition of $k$-tuples is componentwise.) The base case $m=1$ is trivial.

Suppose that $m>1$. By the construction of the sequence, $p_{i}^{n_{\ell_{1}}}$ divides $n_{\ell_{2}}+\ldots+n_{\ell_{m}}$; clearly, $p_{i}^{n_{\ell_{1}}}>n_{\ell_{1}}$ for all $1 \leq i \leq k$. Therefore the Lemma can be applied for $p=p_{i}, k=r=n_{\ell_{1}}$ and $q p^{k}=n_{\ell_{2}}+\ldots+n_{\ell_{m}}$ to obtain

$$
f_{i}\left(n_{\ell_{1}}+n_{\ell_{2}}+\ldots+n_{\ell_{m}}\right)=f_{i}\left(n_{\ell_{1}}\right)+f_{i}\left(n_{\ell_{2}}+\ldots+n_{\ell_{m}}\right) \quad \text { for all } 1 \leq i \leq k
$$

and hence

$$
f\left(n_{\ell_{1}}+n_{\ell_{2}}+\ldots+n_{\ell_{m}}\right)=f\left(n_{\ell_{1}}\right)+f\left(n_{\ell_{2}}+\ldots+n_{\ell_{m}}\right)=f\left(n_{\ell_{1}}\right)+f\left(n_{\ell_{2}}\right)+\ldots+f\left(n_{\ell_{m}}\right)
$$

by the induction hypothesis.
Now consider the values $f\left(n_{1}\right), f\left(n_{2}\right), \ldots$ There exist finitely many possible values of $f$. Hence, there exists an infinite sequence of indices $\ell_{1}<\ell_{2}<\ldots$ such that $f\left(n_{\ell_{1}}\right)=f\left(n_{\ell_{2}}\right)=\ldots$. and thus

$$
f\left(n_{\ell_{m+1}}+n_{\ell_{m+2}}+\ldots+n_{\ell_{m+d}}\right)=f\left(n_{\ell_{m+1}}\right)+\ldots+f\left(n_{\ell_{m+d}}\right)=d \cdot f\left(n_{\ell_{1}}\right)=(\overline{0}, \ldots, \overline{0})
$$

for all $m$. We have found infinitely many suitable numbers.

Solution 2. We use the same Lemma and definition of the function $f$.
Let $S=\{f(n): n \in \mathbb{N}\}$. Obviously, set $S$ is finite. For every $s \in S$ choose the minimal $n_{s}$ such that $f\left(n_{s}\right)=s$. Denote $N=\max _{s \in S} n_{s}$. Moreover, let $g$ be an integer such that $p_{i}^{g}>N$ for each $i=1,2, \ldots, k$. Let $P=\left(p_{1} p_{2} \ldots p_{k}\right)^{g}$.

We claim that

$$
\begin{equation*}
\{f(n) \mid n \in[m P, m P+N]\}=S \tag{1}
\end{equation*}
$$

for every positive integer $m$. In particular, since $(\overline{0}, \ldots, \overline{0})=f(1) \in S$, it follows that for an arbitrary $m$ there exists $n \in[m P, m P+N]$ such that $f(n)=(\overline{0}, \ldots, \overline{0})$. So there are infinitely many suitable numbers.

To prove (1), let $a_{i}=f_{i}(m P)$. Consider all numbers of the form $n_{m, s}=m P+n_{s}$ with $s=\left(s_{1}, \ldots, s_{k}\right) \in S$ (clearly, all $n_{m, s}$ belong to $[m P, m P+N]$ ). Since $n_{s} \leq N<p_{i}^{g}$ and $p_{i}^{g} \mid m P$, we can apply the Lemma for the values $p=p_{i}, r=n_{s}, k=g, q p^{k}=m P$ to obtain

$$
f_{i}\left(n_{m, s}\right)=f_{i}(m P)+f_{i}\left(n_{s}\right)=a_{i}+s_{i} ;
$$

hence for distinct $s, t \in S$ we have $f\left(n_{m, s}\right) \neq f\left(n_{m, t}\right)$.
Thus, the function $f$ attains at least $|S|$ distinct values in $[m P, m P+N]$. Since all these values belong to $S, f$ should attain all possible values in $[m P, m P+N]$.

Comment. Both solutions can be extended to prove the following statements.
Claim 1. For any $K$ there exist infinitely many $n$ divisible by $K$, such that $d \mid \nu_{p_{i}}(n)$ for each $i$.
Claim 2. For any $s \in S$, there exist infinitely many $n \in \mathbb{N}$ such that $f(n)=s$.


# $49^{\text {th }}$ International Mathematical Olympiad Spain 2008 

Shortlisted Problems with Solutions

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## Contributing Countries

Australia, Austria, Belgium, Bulgaria, Canada, Colombia, Croatia, Czech Republic, Estonia, France, Germany, Greece, Hong Kong, India, Iran, Ireland, Japan, Korea (North), Korea (South), Lithuania, Luxembourg, Mexico, Moldova, Netherlands, Pakistan, Peru, Poland, Romania, Russia, Serbia, Slovakia, South Africa, Sweden, Ukraine, United Kingdom, United States of America

## Problem Selection Committee

Vicente Muñoz Velázquez
Juan Manuel Conde Calero
Géza Kós
Marcin Kuczma
Daniel Lasaosa Medarde
Ignasi Mundet i Riera
Svetoslav Savchev

## Algebra

A1. Find all functions $f:(0, \infty) \rightarrow(0, \infty)$ such that

$$
\frac{f(p)^{2}+f(q)^{2}}{f\left(r^{2}\right)+f\left(s^{2}\right)}=\frac{p^{2}+q^{2}}{r^{2}+s^{2}}
$$

for all $p, q, r, s>0$ with $p q=r s$.
Solution. Let $f$ satisfy the given condition. Setting $p=q=r=s=1$ yields $f(1)^{2}=f(1)$ and hence $f(1)=1$. Now take any $x>0$ and set $p=x, q=1, r=s=\sqrt{x}$ to obtain

$$
\frac{f(x)^{2}+1}{2 f(x)}=\frac{x^{2}+1}{2 x} .
$$

This recasts into

$$
\begin{gathered}
x f(x)^{2}+x=x^{2} f(x)+f(x) \\
(x f(x)-1)(f(x)-x)=0
\end{gathered}
$$

And thus,

$$
\begin{equation*}
\text { for every } x>0 \text {, either } f(x)=x \text { or } f(x)=\frac{1}{x} \tag{1}
\end{equation*}
$$

Obviously, if

$$
\begin{equation*}
f(x)=x \quad \text { for all } x>0 \quad \text { or } \quad f(x)=\frac{1}{x} \quad \text { for all } x>0 \tag{2}
\end{equation*}
$$

then the condition of the problem is satisfied. We show that actually these two functions are the only solutions.

So let us assume that there exists a function $f$ satisfying the requirement, other than those in (2). Then $f(a) \neq a$ and $f(b) \neq 1 / b$ for some $a, b>0$. By (1), these values must be $f(a)=1 / a, f(b)=b$. Applying now the equation with $p=a, q=b, r=s=\sqrt{a b}$ we obtain $\left(a^{-2}+b^{2}\right) / 2 f(a b)=\left(a^{2}+b^{2}\right) / 2 a b ;$ equivalently,

$$
\begin{equation*}
f(a b)=\frac{a b\left(a^{-2}+b^{2}\right)}{a^{2}+b^{2}} \tag{3}
\end{equation*}
$$

We know however (see (1)) that $f(a b)$ must be either $a b$ or $1 / a b$. If $f(a b)=a b$ then by (3) $a^{-2}+b^{2}=a^{2}+b^{2}$, so that $a=1$. But, as $f(1)=1$, this contradicts the relation $f(a) \neq a$. Likewise, if $f(a b)=1 / a b$ then (3) gives $a^{2} b^{2}\left(a^{-2}+b^{2}\right)=a^{2}+b^{2}$, whence $b=1$, in contradiction to $f(b) \neq 1 / b$. Thus indeed the functions listed in (2) are the only two solutions.

Comment. The equation has as many as four variables with only one constraint $p q=r s$, leaving three degrees of freedom and providing a lot of information. Various substitutions force various useful properties of the function searched. We sketch one more method to reach conclusion (1); certainly there are many others.

Noticing that $f(1)=1$ and setting, first, $p=q=1, r=\sqrt{x}, s=1 / \sqrt{x}$, and then $p=x, q=1 / x$, $r=s=1$, we obtain two relations, holding for every $x>0$,

$$
\begin{equation*}
f(x)+f\left(\frac{1}{x}\right)=x+\frac{1}{x} \quad \text { and } \quad f(x)^{2}+f\left(\frac{1}{x}\right)^{2}=x^{2}+\frac{1}{x^{2}} . \tag{4}
\end{equation*}
$$

Squaring the first and subtracting the second gives $2 f(x) f(1 / x)=2$. Subtracting this from the second relation of (4) leads to

$$
\left(f(x)-f\left(\frac{1}{x}\right)\right)^{2}=\left(x-\frac{1}{x}\right)^{2} \quad \text { or } \quad f(x)-f\left(\frac{1}{x}\right)= \pm\left(x-\frac{1}{x}\right) .
$$

The last two alternatives combined with the first equation of (4) imply the two alternatives of (1).

A2. (a) Prove the inequality

$$
\frac{x^{2}}{(x-1)^{2}}+\frac{y^{2}}{(y-1)^{2}}+\frac{z^{2}}{(z-1)^{2}} \geq 1
$$

for real numbers $x, y, z \neq 1$ satisfying the condition $x y z=1$.
(b) Show that there are infinitely many triples of rational numbers $x, y, z$ for which this inequality turns into equality.

Solution 1. (a) We start with the substitution

$$
\frac{x}{x-1}=a, \quad \frac{y}{y-1}=b, \quad \frac{z}{z-1}=c, \quad \text { i.e., } \quad x=\frac{a}{a-1}, \quad y=\frac{b}{b-1}, \quad z=\frac{c}{c-1} .
$$

The inequality to be proved reads $a^{2}+b^{2}+c^{2} \geq 1$. The new variables are subject to the constraints $a, b, c \neq 1$ and the following one coming from the condition $x y z=1$,

$$
(a-1)(b-1)(c-1)=a b c .
$$

This is successively equivalent to

$$
\begin{aligned}
a+b+c-1 & =a b+b c+c a \\
2(a+b+c-1) & =(a+b+c)^{2}-\left(a^{2}+b^{2}+c^{2}\right) \\
a^{2}+b^{2}+c^{2}-2 & =(a+b+c)^{2}-2(a+b+c), \\
a^{2}+b^{2}+c^{2}-1 & =(a+b+c-1)^{2}
\end{aligned}
$$

Thus indeed $a^{2}+b^{2}+c^{2} \geq 1$, as desired.
(b) From the equation $a^{2}+b^{2}+c^{2}-1=(a+b+c-1)^{2}$ we see that the proposed inequality becomes an equality if and only if both sums $a^{2}+b^{2}+c^{2}$ and $a+b+c$ have value 1 . The first of them is equal to $(a+b+c)^{2}-2(a b+b c+c a)$. So the instances of equality are described by the system of two equations

$$
a+b+c=1, \quad a b+b c+c a=0
$$

plus the constraint $a, b, c \neq 1$. Elimination of $c$ leads to $a^{2}+a b+b^{2}=a+b$, which we regard as a quadratic equation in $b$,

$$
b^{2}+(a-1) b+a(a-1)=0,
$$

with discriminant

$$
\Delta=(a-1)^{2}-4 a(a-1)=(1-a)(1+3 a) .
$$

We are looking for rational triples $(a, b, c)$; it will suffice to have $a$ rational such that $1-a$ and $1+3 a$ are both squares of rational numbers (then $\Delta$ will be so too). Set $a=k / m$. We want $m-k$ and $m+3 k$ to be squares of integers. This is achieved for instance by taking $m=k^{2}-k+1$ (clearly nonzero); then $m-k=(k-1)^{2}, m+3 k=(k+1)^{2}$. Note that distinct integers $k$ yield distinct values of $a=k / m$.

And thus, if $k$ is any integer and $m=k^{2}-k+1, a=k / m$ then $\Delta=\left(k^{2}-1\right)^{2} / m^{2}$ and the quadratic equation has rational roots $b=\left(m-k \pm k^{2} \mp 1\right) /(2 m)$. Choose e.g. the larger root,

$$
b=\frac{m-k+k^{2}-1}{2 m}=\frac{m+(m-2)}{2 m}=\frac{m-1}{m} .
$$

Computing $c$ from $a+b+c=1$ then gives $c=(1-k) / m$. The condition $a, b, c \neq 1$ eliminates only $k=0$ and $k=1$. Thus, as $k$ varies over integers greater than 1 , we obtain an infinite family of rational triples $(a, b, c)$-and coming back to the original variables $(x=a /(a-1)$ etc.) -an infinite family of rational triples $(x, y, z)$ with the needed property. (A short calculation shows that the resulting triples are $x=-k /(k-1)^{2}, y=k-k^{2}, z=(k-1) / k^{2}$; but the proof was complete without listing them.)

Comment 1. There are many possible variations in handling the equation system $a^{2}+b^{2}+c^{2}=1$, $a+b+c=1(a, b, c \neq 1)$ which of course describes a circle in the ( $a, b, c$ )-space (with three points excluded), and finding infinitely many rational points on it.

Also the initial substitution $x=a /(a-1)$ (etc.) can be successfully replaced by other similar substitutions, e.g. $x=1-1 / \alpha$ (etc.); or $x=x^{\prime}-1$ (etc.); or $1-y z=u$ (etc.)-eventually reducing the inequality to $(\cdots)^{2} \geq 0$, the expression in the parentheses depending on the actual substitution.

Depending on the method chosen, one arrives at various sequences of rational triples $(x, y, z)$ as needed; let us produce just one more such example: $x=(2 r-2) /(r+1)^{2}$, $y=(2 r+2) /(r-1)^{2}$, $z=\left(r^{2}-1\right) / 4$ where $r$ can be any rational number different from 1 or -1 .

Solution 2 (an outline). (a) Without changing variables, just setting $z=1 / x y$ and clearing fractions, the proposed inequality takes the form

$$
(x y-1)^{2}\left(x^{2}(y-1)^{2}+y^{2}(x-1)^{2}\right)+(x-1)^{2}(y-1)^{2} \geq(x-1)^{2}(y-1)^{2}(x y-1)^{2} .
$$

With the notation $p=x+y, q=x y$ this becomes, after lengthy routine manipulation and a lot of cancellation

$$
q^{4}-6 q^{3}+2 p q^{2}+9 q^{2}-6 p q+p^{2} \geq 0
$$

It is not hard to notice that the expression on the left is just $\left(q^{2}-3 q+p\right)^{2}$, hence nonnegative.
(Without introducing $p$ and $q$, one is of course led with some more work to the same expression, just written in terms of $x$ and $y$; but then it is not that easy to see that it is a square.)
(b) To have equality, one needs $q^{2}-3 q+p=0$. Note that $x$ and $y$ are the roots of the quadratic trinomial (in a formal variable $t$ ): $t^{2}-p t+q$. When $q^{2}-3 q+p=0$, the discriminant equals

$$
\delta=p^{2}-4 q=\left(3 q-q^{2}\right)^{2}-4 q=q(q-1)^{2}(q-4)
$$

Now it suffices to have both $q$ and $q-4$ squares of rational numbers (then $p=3 q-q^{2}$ and $\sqrt{\delta}$ are also rational, and so are the roots of the trinomial). On setting $q=(n / m)^{2}=4+(l / m)^{2}$ the requirement becomes $4 m^{2}+l^{2}=n^{2}$ (with $l, m, n$ being integers). This is just the Pythagorean equation, known to have infinitely many integer solutions.

Comment 2. Part (a) alone might also be considered as a possible contest problem (in the category of easy problems).

A3. Let $S \subseteq \mathbb{R}$ be a set of real numbers. We say that a pair $(f, g)$ of functions from $S$ into $S$ is a Spanish Couple on $S$, if they satisfy the following conditions:
(i) Both functions are strictly increasing, i.e. $f(x)<f(y)$ and $g(x)<g(y)$ for all $x, y \in S$ with $x<y$;
(ii) The inequality $f(g(g(x)))<g(f(x))$ holds for all $x \in S$.

Decide whether there exists a Spanish Couple
(a) on the set $S=\mathbb{N}$ of positive integers;
(b) on the set $S=\{a-1 / b: a, b \in \mathbb{N}\}$.

Solution. We show that the answer is NO for part (a), and YES for part (b).
(a) Throughout the solution, we will use the notation $g_{k}(x)=\overbrace{g(g(\ldots g}^{k}(x) \ldots))$, including $g_{0}(x)=x$ as well.

Suppose that there exists a Spanish Couple $(f, g)$ on the set $\mathbb{N}$. From property (i) we have $f(x) \geq x$ and $g(x) \geq x$ for all $x \in \mathbb{N}$.

We claim that $g_{k}(x) \leq f(x)$ for all $k \geq 0$ and all positive integers $x$. The proof is done by induction on $k$. We already have the base case $k=0$ since $x \leq f(x)$. For the induction step from $k$ to $k+1$, apply the induction hypothesis on $g_{2}(x)$ instead of $x$, then apply (ii):

$$
g\left(g_{k+1}(x)\right)=g_{k}\left(g_{2}(x)\right) \leq f\left(g_{2}(x)\right)<g(f(x))
$$

Since $g$ is increasing, it follows that $g_{k+1}(x)<f(x)$. The claim is proven.
If $g(x)=x$ for all $x \in \mathbb{N}$ then $f(g(g(x)))=f(x)=g(f(x))$, and we have a contradiction with (ii). Therefore one can choose an $x_{0} \in S$ for which $x_{0}<g\left(x_{0}\right)$. Now consider the sequence $x_{0}, x_{1}, \ldots$ where $x_{k}=g_{k}\left(x_{0}\right)$. The sequence is increasing. Indeed, we have $x_{0}<g\left(x_{0}\right)=x_{1}$, and $x_{k}<x_{k+1}$ implies $x_{k+1}=g\left(x_{k}\right)<g\left(x_{k+1}\right)=x_{k+2}$.

Hence, we obtain a strictly increasing sequence $x_{0}<x_{1}<\ldots$ of positive integers which on the other hand has an upper bound, namely $f\left(x_{0}\right)$. This cannot happen in the set $\mathbb{N}$ of positive integers, thus no Spanish Couple exists on $\mathbb{N}$.
(b) We present a Spanish Couple on the set $S=\{a-1 / b: a, b \in \mathbb{N}\}$.

Let

$$
\begin{aligned}
f(a-1 / b) & =a+1-1 / b, \\
g(a-1 / b) & =a-1 /\left(b+3^{a}\right) .
\end{aligned}
$$

These functions are clearly increasing. Condition (ii) holds, since

$$
f(g(g(a-1 / b)))=(a+1)-1 /\left(b+2 \cdot 3^{a}\right)<(a+1)-1 /\left(b+3^{a+1}\right)=g(f(a-1 / b)) .
$$

Comment. Another example of a Spanish couple is $f(a-1 / b)=3 a-1 / b, g(a-1 / b)=a-1 /(a+b)$. More generally, postulating $f(a-1 / b)=h(a)-1 / b, \quad g(a-1 / b)=a-1 / G(a, b)$ with $h$ increasing and $G$ increasing in both variables, we get that $f \circ g \circ g<g \circ f$ holds if $G(a, G(a, b))<G(h(a), b)$. A search just among linear functions $h(a)=C a, G(a, b)=A a+B b$ results in finding that any integers $A>0, C>2$ and $B=1$ produce a Spanish couple (in the example above, $A=1, C=3$ ). The proposer's example results from taking $h(a)=a+1, G(a, b)=3^{a}+b$.

A4. For an integer $m$, denote by $t(m)$ the unique number in $\{1,2,3\}$ such that $m+t(m)$ is a multiple of 3. A function $f: \mathbb{Z} \rightarrow \mathbb{Z}$ satisfies $f(-1)=0, f(0)=1, f(1)=-1$ and

$$
f\left(2^{n}+m\right)=f\left(2^{n}-t(m)\right)-f(m) \text { for all integers } m, n \geq 0 \text { with } 2^{n}>m
$$

Prove that $f(3 p) \geq 0$ holds for all integers $p \geq 0$.
Solution. The given conditions determine $f$ uniquely on the positive integers. The signs of $f(1), f(2), \ldots$ seem to change quite erratically. However values of the form $f\left(2^{n}-t(m)\right)$ are sufficient to compute directly any functional value. Indeed, let $n>0$ have base 2 representation $n=2^{a_{0}}+2^{a_{1}}+\cdots+2^{a_{k}}, a_{0}>a_{1}>\cdots>a_{k} \geq 0$, and let $n_{j}=2^{a_{j}}+2^{a_{j-1}}+\cdots+2^{a_{k}}, j=0, \ldots, k$. Repeated applications of the recurrence show that $f(n)$ is an alternating sum of the quantities $f\left(2^{a_{j}}-t\left(n_{j+1}\right)\right)$ plus $(-1)^{k+1}$. (The exact formula is not needed for our proof.)

So we focus attention on the values $f\left(2^{n}-1\right), f\left(2^{n}-2\right)$ and $f\left(2^{n}-3\right)$. Six cases arise; more specifically,
$t\left(2^{2 k}-3\right)=2, t\left(2^{2 k}-2\right)=1, t\left(2^{2 k}-1\right)=3, t\left(2^{2 k+1}-3\right)=1, t\left(2^{2 k+1}-2\right)=3, t\left(2^{2 k+1}-1\right)=2$.
Claim. For all integers $k \geq 0$ the following equalities hold:

$$
\begin{array}{lll}
f\left(2^{2 k+1}-3\right)=0, & f\left(2^{2 k+1}-2\right)=3^{k}, & f\left(2^{2 k+1}-1\right)=-3^{k} \\
f\left(2^{2 k+2}-3\right)=-3^{k}, & f\left(2^{2 k+2}-2\right)=-3^{k}, & f\left(2^{2 k+2}-1\right)=2 \cdot 3^{k} .
\end{array}
$$

Proof. By induction on $k$. The base $k=0$ comes down to checking that $f(2)=-1$ and $f(3)=2$; the given values $f(-1)=0, f(0)=1, f(1)=-1$ are also needed. Suppose the claim holds for $k-1$. For $f\left(2^{2 k+1}-t(m)\right)$, the recurrence formula and the induction hypothesis yield

$$
\begin{aligned}
& f\left(2^{2 k+1}-3\right)=f\left(2^{2 k}+\left(2^{2 k}-3\right)\right)=f\left(2^{2 k}-2\right)-f\left(2^{2 k}-3\right)=-3^{k-1}+3^{k-1}=0, \\
& f\left(2^{2 k+1}-2\right)=f\left(2^{2 k}+\left(2^{2 k}-2\right)\right)=f\left(2^{2 k}-1\right)-f\left(2^{2 k}-2\right)=2 \cdot 3^{k-1}+3^{k-1}=3^{k}, \\
& f\left(2^{2 k+1}-1\right)=f\left(2^{2 k}+\left(2^{2 k}-1\right)\right)=f\left(2^{2 k}-3\right)-f\left(2^{2 k}-1\right)=-3^{k-1}-2 \cdot 3^{k-1}=-3^{k} .
\end{aligned}
$$

For $f\left(2^{2 k+2}-t(m)\right)$ we use the three equalities just established:

$$
\begin{aligned}
& f\left(2^{2 k+2}-3\right)=f\left(2^{2 k+1}+\left(2^{2 k+1}-3\right)\right)=f\left(2^{2 k+1}-1\right)-f\left(2^{2 k+1}-3\right)=-3^{k}-0=-3^{k} \\
& f\left(2^{2 k+2}-2\right)=f\left(2^{2 k+1}+\left(2^{2 k+1}-2\right)\right)=f\left(2^{2 k+1}-3\right)-f\left(2^{2 k}-2\right)=0-3^{k}=-3^{k} \\
& f\left(2^{2 k+2}-1\right)=f\left(2^{2 k+1}+\left(2^{2 k+1}-1\right)\right)=f\left(2^{2 k+1}-2\right)-f\left(2^{2 k+1}-1\right)=3^{k}+3^{k}=2 \cdot 3^{k}
\end{aligned}
$$

The claim follows.
A closer look at the six cases shows that $f\left(2^{n}-t(m)\right) \geq 3^{(n-1) / 2}$ if $2^{n}-t(m)$ is divisible by 3 , and $f\left(2^{n}-t(m)\right) \leq 0$ otherwise. On the other hand, note that $2^{n}-t(m)$ is divisible by 3 if and only if $2^{n}+m$ is. Therefore, for all nonnegative integers $m$ and $n$,
(i) $f\left(2^{n}-t(m)\right) \geq 3^{(n-1) / 2}$ if $2^{n}+m$ is divisible by 3 ;
(ii) $f\left(2^{n}-t(m)\right) \leq 0$ if $2^{n}+m$ is not divisible by 3 .

One more (direct) consequence of the claim is that $\left|f\left(2^{n}-t(m)\right)\right| \leq \frac{2}{3} \cdot 3^{n / 2}$ for all $m, n \geq 0$.
The last inequality enables us to find an upper bound for $|f(m)|$ for $m$ less than a given power of 2 . We prove by induction on $n$ that $|f(m)| \leq 3^{n / 2}$ holds true for all integers $m, n \geq 0$ with $2^{n}>m$.

The base $n=0$ is clear as $f(0)=1$. For the inductive step from $n$ to $n+1$, let $m$ and $n$ satisfy $2^{n+1}>m$. If $m<2^{n}$, we are done by the inductive hypothesis. If $m \geq 2^{n}$ then $m=2^{n}+k$ where $2^{n}>k \geq 0$. Now, by $\left|f\left(2^{n}-t(k)\right)\right| \leq \frac{2}{3} \cdot 3^{n / 2}$ and the inductive assumption,

$$
|f(m)|=\left|f\left(2^{n}-t(k)\right)-f(k)\right| \leq\left|f\left(2^{n}-t(k)\right)\right|+|f(k)| \leq \frac{2}{3} \cdot 3^{n / 2}+3^{n / 2}<3^{(n+1) / 2}
$$

The induction is complete.
We proceed to prove that $f(3 p) \geq 0$ for all integers $p \geq 0$. Since $3 p$ is not a power of 2 , its binary expansion contains at least two summands. Hence one can write $3 p=2^{a}+2^{b}+c$ where $a>b$ and $2^{b}>c \geq 0$. Applying the recurrence formula twice yields

$$
f(3 p)=f\left(2^{a}+2^{b}+c\right)=f\left(2^{a}-t\left(2^{b}+c\right)\right)-f\left(2^{b}-t(c)\right)+f(c) .
$$

Since $2^{a}+2^{b}+c$ is divisible by 3 , we have $f\left(2^{a}-t\left(2^{b}+c\right)\right) \geq 3^{(a-1) / 2}$ by (i). Since $2^{b}+c$ is not divisible by 3 , we have $f\left(2^{b}-t(c)\right) \leq 0$ by (ii). Finally $|f(c)| \leq 3^{b / 2}$ as $2^{b}>c \geq 0$, so that $f(c) \geq-3^{b / 2}$. Therefore $f(3 p) \geq 3^{(a-1) / 2}-3^{b / 2}$ which is nonnegative because $a>b$.

A5. Let $a, b, c, d$ be positive real numbers such that

$$
a b c d=1 \quad \text { and } \quad a+b+c+d>\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a} .
$$

Prove that

$$
a+b+c+d<\frac{b}{a}+\frac{c}{b}+\frac{d}{c}+\frac{a}{d}
$$

Solution. We show that if $a b c d=1$, the sum $a+b+c+d$ cannot exceed a certain weighted mean of the expressions $\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a}$ and $\frac{b}{a}+\frac{c}{b}+\frac{d}{c}+\frac{a}{d}$.

By applying the AM-GM inequality to the numbers $\frac{a}{b}, \frac{a}{b}, \frac{b}{c}$ and $\frac{a}{d}$, we obtain

$$
a=\sqrt[4]{\frac{a^{4}}{a b c d}}=\sqrt[4]{\frac{a}{b} \cdot \frac{a}{b} \cdot \frac{b}{c} \cdot \frac{a}{d}} \leq \frac{1}{4}\left(\frac{a}{b}+\frac{a}{b}+\frac{b}{c}+\frac{a}{d}\right)
$$

Analogously,

$$
b \leq \frac{1}{4}\left(\frac{b}{c}+\frac{b}{c}+\frac{c}{d}+\frac{b}{a}\right), \quad c \leq \frac{1}{4}\left(\frac{c}{d}+\frac{c}{d}+\frac{d}{a}+\frac{c}{b}\right) \quad \text { and } \quad d \leq \frac{1}{4}\left(\frac{d}{a}+\frac{d}{a}+\frac{a}{b}+\frac{d}{c}\right) .
$$

Summing up these estimates yields

$$
a+b+c+d \leq \frac{3}{4}\left(\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a}\right)+\frac{1}{4}\left(\frac{b}{a}+\frac{c}{b}+\frac{d}{c}+\frac{a}{d}\right) .
$$

In particular, if $a+b+c+d>\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a}$ then $a+b+c+d<\frac{b}{a}+\frac{c}{b}+\frac{d}{c}+\frac{a}{d}$.
Comment. The estimate in the above solution was obtained by applying the AM-GM inequality to each column of the $4 \times 4$ array

$$
\begin{array}{llll}
a / b & b / c & c / d & d / a \\
a / b & b / c & c / d & d / a \\
b / c & c / d & d / a & a / b \\
a / d & b / a & c / b & d / c
\end{array}
$$

and adding up the resulting inequalities. The same table yields a stronger bound: If $a, b, c, d>0$ and $a b c d=1$ then

$$
\left(\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a}\right)^{3}\left(\frac{b}{a}+\frac{c}{b}+\frac{d}{c}+\frac{a}{d}\right) \geq(a+b+c+d)^{4}
$$

It suffices to apply Hölder's inequality to the sequences in the four rows, with weights $1 / 4$ :

$$
\begin{gathered}
\left(\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a}\right)^{1 / 4}\left(\frac{a}{b}+\frac{b}{c}+\frac{c}{d}+\frac{d}{a}\right)^{1 / 4}\left(\frac{b}{c}+\frac{c}{d}+\frac{d}{a}+\frac{a}{b}\right)^{1 / 4}\left(\frac{a}{d}+\frac{b}{a}+\frac{c}{b}+\frac{d}{c}\right)^{1 / 4} \\
\geq\left(\frac{a a b a}{b b c d}\right)^{1 / 4}+\left(\frac{b b c b}{c c d a}\right)^{1 / 4}+\left(\frac{c c d c}{d d a b}\right)^{1 / 4}+\left(\frac{d d a d}{a a b c}\right)^{1 / 4}=a+b+c+d
\end{gathered}
$$

A6. Let $f: \mathbb{R} \rightarrow \mathbb{N}$ be a function which satisfies

$$
\begin{equation*}
f\left(x+\frac{1}{f(y)}\right)=f\left(y+\frac{1}{f(x)}\right) \quad \text { for all } x, y \in \mathbb{R} \tag{1}
\end{equation*}
$$

Prove that there is a positive integer which is not a value of $f$.
Solution. Suppose that the statement is false and $f(\mathbb{R})=\mathbb{N}$. We prove several properties of the function $f$ in order to reach a contradiction.

To start with, observe that one can assume $f(0)=1$. Indeed, let $a \in \mathbb{R}$ be such that $f(a)=1$, and consider the function $g(x)=f(x+a)$. By substituting $x+a$ and $y+a$ for $x$ and $y$ in (1), we have

$$
g\left(x+\frac{1}{g(y)}\right)=f\left(x+a+\frac{1}{f(y+a)}\right)=f\left(y+a+\frac{1}{f(x+a)}\right)=g\left(y+\frac{1}{g(x)}\right)
$$

So $g$ satisfies the functional equation (1), with the additional property $g(0)=1$. Also, $g$ and $f$ have the same set of values: $g(\mathbb{R})=f(\mathbb{R})=\mathbb{N}$. Henceforth we assume $f(0)=1$.
Claim 1. For an arbitrary fixed $c \in \mathbb{R}$ we have $\left\{f\left(c+\frac{1}{n}\right): n \in \mathbb{N}\right\}=\mathbb{N}$.
Proof. Equation (1) and $f(\mathbb{R})=\mathbb{N}$ imply
$f(\mathbb{R})=\left\{f\left(x+\frac{1}{f(c)}\right): x \in \mathbb{R}\right\}=\left\{f\left(c+\frac{1}{f(x)}\right): x \in \mathbb{R}\right\} \subset\left\{f\left(c+\frac{1}{n}\right): n \in \mathbb{N}\right\} \subset f(\mathbb{R})$.
The claim follows.
We will use Claim 1 in the special cases $c=0$ and $c=1 / 3$ :

$$
\begin{equation*}
\left\{f\left(\frac{1}{n}\right): n \in \mathbb{N}\right\}=\left\{f\left(\frac{1}{3}+\frac{1}{n}\right): n \in \mathbb{N}\right\}=\mathbb{N} \tag{2}
\end{equation*}
$$

Claim 2. If $f(u)=f(v)$ for some $u, v \in \mathbb{R}$ then $f(u+q)=f(v+q)$ for all nonnegative rational $q$. Furthermore, if $f(q)=1$ for some nonnegative rational $q$ then $f(k q)=1$ for all $k \in \mathbb{N}$.
Proof. For all $x \in \mathbb{R}$ we have by (1)

$$
f\left(u+\frac{1}{f(x)}\right)=f\left(x+\frac{1}{f(u)}\right)=f\left(x+\frac{1}{f(v)}\right)=f\left(v+\frac{1}{f(x)}\right) .
$$

Since $f(x)$ attains all positive integer values, this yields $f(u+1 / n)=f(v+1 / n)$ for all $n \in \mathbb{N}$. Let $q=k / n$ be a positive rational number. Then $k$ repetitions of the last step yield

$$
f(u+q)=f\left(u+\frac{k}{n}\right)=f\left(v+\frac{k}{n}\right)=f(v+q) .
$$

Now let $f(q)=1$ for some nonnegative rational $q$, and let $k \in \mathbb{N}$. As $f(0)=1$, the previous conclusion yields successively $f(q)=f(2 q), f(2 q)=f(3 q), \ldots, f((k-1) q)=f(k q)$, as needed.
Claim 3. The equality $f(q)=f(q+1)$ holds for all nonnegative rational $q$.
Proof. Let $m$ be a positive integer such that $f(1 / m)=1$. Such an $m$ exists by (2). Applying the second statement of Claim 2 with $q=1 / m$ and $k=m$ yields $f(1)=1$.

Given that $f(0)=f(1)=1$, the first statement of Claim 2 implies $f(q)=f(q+1)$ for all nonnegative rational $q$.

Claim 4. The equality $f\left(\frac{1}{n}\right)=n$ holds for every $n \in \mathbb{N}$.
Proof. For a nonnegative rational $q$ we set $x=q, y=0$ in (1) and use Claim 3 to obtain

$$
f\left(\frac{1}{f(q)}\right)=f\left(q+\frac{1}{f(0)}\right)=f(q+1)=f(q)
$$

By (2), for each $n \in \mathbb{N}$ there exists a $k \in \mathbb{N}$ such that $f(1 / k)=n$. Applying the last equation with $q=1 / k$, we have

$$
n=f\left(\frac{1}{k}\right)=f\left(\frac{1}{f(1 / k)}\right)=f\left(\frac{1}{n}\right) .
$$

Now we are ready to obtain a contradiction. Let $n \in \mathbb{N}$ be such that $f(1 / 3+1 / n)=1$. Such an $n$ exists by (2). Let $1 / 3+1 / n=s / t$, where $s, t \in \mathbb{N}$ are coprime. Observe that $t>1$ as $1 / 3+1 / n$ is not an integer. Choose $k, l \in \mathbb{N}$ so that that $k s-l t=1$.

Because $f(0)=f(s / t)=1$, Claim 2 implies $f(k s / t)=1$. Now $f(k s / t)=f(1 / t+l)$; on the other hand $f(1 / t+l)=f(1 / t)$ by $l$ successive applications of Claim 3. Finally, $f(1 / t)=t$ by Claim 4, leading to the impossible $t=1$. The solution is complete.

A7. Prove that for any four positive real numbers $a, b, c, d$ the inequality

$$
\frac{(a-b)(a-c)}{a+b+c}+\frac{(b-c)(b-d)}{b+c+d}+\frac{(c-d)(c-a)}{c+d+a}+\frac{(d-a)(d-b)}{d+a+b} \geq 0
$$

holds. Determine all cases of equality.
Solution 1. Denote the four terms by

$$
A=\frac{(a-b)(a-c)}{a+b+c}, \quad B=\frac{(b-c)(b-d)}{b+c+d}, \quad C=\frac{(c-d)(c-a)}{c+d+a}, \quad D=\frac{(d-a)(d-b)}{d+a+b} .
$$

The expression $2 A$ splits into two summands as follows,

$$
2 A=A^{\prime}+A^{\prime \prime} \quad \text { where } \quad A^{\prime}=\frac{(a-c)^{2}}{a+b+c}, \quad A^{\prime \prime}=\frac{(a-c)(a-2 b+c)}{a+b+c}
$$

this is easily verified. We analogously represent $2 B=B^{\prime}+B^{\prime \prime}, 2 C=C^{\prime}+C^{\prime \prime}, 2 B=D^{\prime}+D^{\prime \prime}$ and examine each of the sums $A^{\prime}+B^{\prime}+C^{\prime}+D^{\prime}$ and $A^{\prime \prime}+B^{\prime \prime}+C^{\prime \prime}+D^{\prime \prime}$ separately.

Write $s=a+b+c+d$; the denominators become $s-d, s-a, s-b, s-c$. By the CauchySchwarz inequality,

$$
\begin{aligned}
& \left(\frac{|a-c|}{\sqrt{s-d}} \cdot \sqrt{s-d}+\frac{|b-d|}{\sqrt{s-a}} \cdot \sqrt{s-a}+\frac{|c-a|}{\sqrt{s-b}} \cdot \sqrt{s-b}+\frac{|d-b|}{\sqrt{s-c}} \cdot \sqrt{s-c}\right)^{2} \\
& \quad \leq\left(\frac{(a-c)^{2}}{s-d}+\frac{(b-d)^{2}}{s-a}+\frac{(c-a)^{2}}{s-b}+\frac{(d-b)^{2}}{s-c}\right)(4 s-s)=3 s\left(A^{\prime}+B^{\prime}+C^{\prime}+D^{\prime}\right)
\end{aligned}
$$

Hence

$$
\begin{equation*}
A^{\prime}+B^{\prime}+C^{\prime}+D^{\prime} \geq \frac{(2|a-c|+2|b-d|)^{2}}{3 s} \geq \frac{16 \cdot|a-c| \cdot|b-d|}{3 s} \tag{1}
\end{equation*}
$$

Next we estimate the absolute value of the other sum. We couple $A^{\prime \prime}$ with $C^{\prime \prime}$ to obtain

$$
\begin{aligned}
A^{\prime \prime}+C^{\prime \prime} & =\frac{(a-c)(a+c-2 b)}{s-d}+\frac{(c-a)(c+a-2 d)}{s-b} \\
& =\frac{(a-c)(a+c-2 b)(s-b)+(c-a)(c+a-2 d)(s-d)}{(s-d)(s-b)} \\
& =\frac{(a-c)(-2 b(s-b)-b(a+c)+2 d(s-d)+d(a+c))}{s(a+c)+b d} \\
& =\frac{3(a-c)(d-b)(a+c)}{M}, \quad \text { with } \quad M=s(a+c)+b d .
\end{aligned}
$$

Hence by cyclic shift

$$
B^{\prime \prime}+D^{\prime \prime}=\frac{3(b-d)(a-c)(b+d)}{N}, \quad \text { with } \quad N=s(b+d)+c a .
$$

Thus

$$
\begin{equation*}
A^{\prime \prime}+B^{\prime \prime}+C^{\prime \prime}+D^{\prime \prime}=3(a-c)(b-d)\left(\frac{b+d}{N}-\frac{a+c}{M}\right)=\frac{3(a-c)(b-d) W}{M N} \tag{2}
\end{equation*}
$$

where

$$
\begin{equation*}
W=(b+d) M-(a+c) N=b d(b+d)-a c(a+c) . \tag{3}
\end{equation*}
$$

Note that

$$
\begin{equation*}
M N>(a c(a+c)+b d(b+d)) s \geq|W| \cdot s \tag{4}
\end{equation*}
$$

Now (2) and (4) yield

$$
\begin{equation*}
\left|A^{\prime \prime}+B^{\prime \prime}+C^{\prime \prime}+D^{\prime \prime}\right| \leq \frac{3 \cdot|a-c| \cdot|b-d|}{s} \tag{5}
\end{equation*}
$$

Combined with (1) this results in

$$
\begin{aligned}
& 2(A+B+C+D)=\left(A^{\prime}+B^{\prime}+C^{\prime}+D^{\prime}\right)+\left(A^{\prime \prime}+B^{\prime \prime}+C^{\prime \prime}+D^{\prime \prime}\right) \\
& \quad \geq \frac{16 \cdot|a-c| \cdot|b-d|}{3 s}-\frac{3 \cdot|a-c| \cdot|b-d|}{s}=\frac{7 \cdot|a-c| \cdot|b-d|}{3(a+b+c+d)} \geq 0
\end{aligned}
$$

This is the required inequality. From the last line we see that equality can be achieved only if either $a=c$ or $b=d$. Since we also need equality in (1), this implies that actually $a=c$ and $b=d$ must hold simultaneously, which is obviously also a sufficient condition.

Solution 2. We keep the notations $A, B, C, D, s$, and also $M, N, W$ from the preceding solution; the definitions of $M, N, W$ and relations (3), (4) in that solution did not depend on the foregoing considerations. Starting from

$$
2 A=\frac{(a-c)^{2}+3(a+c)(a-c)}{a+b+c}-2 a+2 c
$$

we get

$$
\begin{aligned}
2(A & +C)=(a-c)^{2}\left(\frac{1}{s-d}+\frac{1}{s-b}\right)+3(a+c)(a-c)\left(\frac{1}{s-d}-\frac{1}{s-b}\right) \\
& =(a-c)^{2} \frac{2 s-b-d}{M}+3(a+c)(a-c) \cdot \frac{d-b}{M}=\frac{p(a-c)^{2}-3(a+c)(a-c)(b-d)}{M}
\end{aligned}
$$

where $p=2 s-b-d=s+a+c$. Similarly, writing $q=s+b+d$ we have

$$
2(B+D)=\frac{q(b-d)^{2}-3(b+d)(b-d)(c-a)}{N} ;
$$

specific grouping of terms in the numerators has its aim. Note that $p q>2 s^{2}$. By adding the fractions expressing $2(A+C)$ and $2(B+D)$,

$$
2(A+B+C+D)=\frac{p(a-c)^{2}}{M}+\frac{3(a-c)(b-d) W}{M N}+\frac{q(b-d)^{2}}{N}
$$

with $W$ defined by (3).
Substitution $x=(a-c) / M, y=(b-d) / N$ brings the required inequality to the form

$$
\begin{equation*}
2(A+B+C+D)=M p x^{2}+3 W x y+N q y^{2} \geq 0 \tag{6}
\end{equation*}
$$

It will be enough to verify that the discriminant $\Delta=9 W^{2}-4 M N p q$ of the quadratic trinomial $M p t^{2}+3 W t+N q$ is negative; on setting $t=x / y$ one then gets (6). The first inequality in (4) together with $p q>2 s^{2}$ imply $4 M N p q>8 s^{3}(a c(a+c)+b d(b+d))$. Since

$$
(a+c) s^{3}>(a+c)^{4} \geq 4 a c(a+c)^{2} \quad \text { and likewise } \quad(b+d) s^{3}>4 b d(b+d)^{2}
$$

the estimate continues as follows,

$$
4 M N p q>8\left(4(a c)^{2}(a+c)^{2}+4(b d)^{2}(b+d)^{2}\right)>32(b d(b+d)-a c(a+c))^{2}=32 W^{2} \geq 9 W^{2}
$$

Thus indeed $\Delta<0$. The desired inequality (6) hence results. It becomes an equality if and only if $x=y=0$; equivalently, if and only if $a=c$ and simultaneously $b=d$.

Comment. The two solutions presented above do not differ significantly; large portions overlap. The properties of the number $W$ turn out to be crucial in both approaches. The Cauchy-Schwarz inequality, applied in the first solution, is avoided in the second, which requires no knowledge beyond quadratic trinomials.

The estimates in the proof of $\Delta<0$ in the second solution seem to be very wasteful. However, they come close to sharp when the terms in one of the pairs $(a, c),(b, d)$ are equal and much bigger than those in the other pair.

In attempts to prove the inequality by just considering the six cases of arrangement of the numbers $a, b, c, d$ on the real line, one soon discovers that the cases which create real trouble are precisely those in which $a$ and $c$ are both greater or both smaller than $b$ and $d$.

## Solution 3.

$$
\begin{gathered}
(a-b)(a-c)(a+b+d)(a+c+d)(b+c+d)= \\
=((a-b)(a+b+d))((a-c)(a+c+d))(b+c+d)= \\
=\left(a^{2}+a d-b^{2}-b d\right)\left(a^{2}+a d-c^{2}-c d\right)(b+c+d)= \\
=\left(a^{4}+2 a^{3} d-a^{2} b^{2}-a^{2} b d-a^{2} c^{2}-a^{2} c d+a^{2} d^{2}-a b^{2} d-a b d^{2}-a c^{2} d-a c d^{2}+b^{2} c^{2}+b^{2} c d+b c^{2} d+b c d^{2}\right)(b+c+d)= \\
=a^{4} b+a^{4} c+a^{4} d+\left(b^{3} c^{2}+a^{2} d^{3}\right)-a^{2} c^{3}+\left(2 a^{3} d^{2}-b^{3} a^{2}+c^{3} b^{2}\right)+ \\
+\left(b^{3} c d-c^{3} d a-d^{3} a b\right)+\left(2 a^{3} b d+c^{3} d b-d^{3} a c\right)+\left(2 a^{3} c d-b^{3} d a+d^{3} b c\right) \\
+\left(-a^{2} b^{2} c+3 b^{2} c^{2} d-2 a c^{2} d^{2}\right)+\left(-2 a^{2} b^{2} d+2 b c^{2} d^{2}\right)+\left(-a^{2} b c^{2}-2 a^{2} c^{2} d-2 a b^{2} d^{2}+2 b^{2} c d^{2}\right)+ \\
+\left(-2 a^{2} b c d-a b^{2} c d-a b c^{2} d-2 a b c d^{2}\right)
\end{gathered}
$$

Introducing the notation $S_{x y z w}=\sum_{c y c} a^{x} b^{y} c^{z} d^{w}$, one can write

$$
\begin{gathered}
\sum_{c y c}(a-b)(a-c)(a+b+d)(a+c+d)(b+c+d)= \\
=S_{4100}+S_{4010}+S_{4001}+2 S_{3200}-S_{3020}+2 S_{3002}-S_{3110}+2 S_{3101}+2 S_{3011}-3 S_{2120}-6 S_{2111}= \\
+\left(S_{4100}+S_{4001}+\frac{1}{2} S_{3110}+\frac{1}{2} S_{3011}-3 S_{2120}\right)+ \\
+\left(S_{4010}-S_{3020}-\frac{3}{2} S_{3110}+\frac{3}{2} S_{3011}+\frac{9}{16} S_{2210}+\frac{9}{16} S_{2201}-\frac{9}{8} S_{2111}\right)+ \\
+\frac{9}{16}\left(S_{3200}-S_{2210}-S_{2201}+S_{3002}\right)+\frac{23}{16}\left(S_{3200}-2 S_{3101}+S_{3002}\right)+\frac{39}{8}\left(S_{3101}-S_{2111}\right),
\end{gathered}
$$

where the expressions

$$
\begin{gathered}
S_{4100}+S_{4001}+\frac{1}{2} S_{3110}+\frac{1}{2} S_{3011}-3 S_{2120}=\sum_{c y c}\left(a^{4} b+b c^{4}+\frac{1}{2} a^{3} b c+\frac{1}{2} a b c^{3}-3 a^{2} b c^{2}\right), \\
S_{4010}-S_{3020}-\frac{3}{2} S_{3110}+\frac{3}{2} S_{3011}+\frac{9}{16} S_{2210}+\frac{9}{16} S_{2201}-\frac{9}{8} S_{2111}=\sum_{c y c} a^{2} c\left(a-c-\frac{3}{4} b+\frac{3}{4} d\right)^{2}, \\
S_{3200}-S_{2210}-S_{2201}+S_{3002}=\sum_{c y c} b^{2}\left(a^{3}-a^{2} c-a c^{2}+c^{3}\right)=\sum_{c y c} b^{2}(a+c)(a-c)^{2},
\end{gathered}
$$

$$
S_{3200}-2 S_{3101}+S_{3002}=\sum_{c y c} a^{3}(b-d)^{2} \quad \text { and } \quad S_{3101}-S_{2111}=\frac{1}{3} \sum_{c y c} b d\left(2 a^{3}+c^{3}-3 a^{2} c\right)
$$

are all nonnegative.

## Combinatorics

C1. In the plane we consider rectangles whose sides are parallel to the coordinate axes and have positive length. Such a rectangle will be called a box. Two boxes intersect if they have a common point in their interior or on their boundary.

Find the largest $n$ for which there exist $n$ boxes $B_{1}, \ldots, B_{n}$ such that $B_{i}$ and $B_{j}$ intersect if and only if $i \not \equiv j \pm 1(\bmod n)$.

Solution. The maximum number of such boxes is 6 . One example is shown in the figure.


Now we show that 6 is the maximum. Suppose that boxes $B_{1}, \ldots, B_{n}$ satisfy the condition. Let the closed intervals $I_{k}$ and $J_{k}$ be the projections of $B_{k}$ onto the $x$ - and $y$-axis, for $1 \leq k \leq n$.

If $B_{i}$ and $B_{j}$ intersect, with a common point $(x, y)$, then $x \in I_{i} \cap I_{j}$ and $y \in J_{i} \cap J_{j}$. So the intersections $I_{i} \cap I_{j}$ and $J_{i} \cap J_{j}$ are nonempty. Conversely, if $x \in I_{i} \cap I_{j}$ and $y \in J_{i} \cap J_{j}$ for some real numbers $x, y$, then $(x, y)$ is a common point of $B_{i}$ and $B_{j}$. Putting it around, $B_{i}$ and $B_{j}$ are disjoint if and only if their projections on at least one coordinate axis are disjoint.

For brevity we call two boxes or intervals adjacent if their indices differ by 1 modulo $n$, and nonadjacent otherwise.

The adjacent boxes $B_{k}$ and $B_{k+1}$ do not intersect for each $k=1, \ldots, n$. Hence $\left(I_{k}, I_{k+1}\right)$ or ( $J_{k}, J_{k+1}$ ) is a pair of disjoint intervals, $1 \leq k \leq n$. So there are at least $n$ pairs of disjoint intervals among $\left(I_{1}, I_{2}\right), \ldots,\left(I_{n-1}, I_{n}\right),\left(I_{n}, I_{1}\right) ;\left(J_{1}, J_{2}\right), \ldots,\left(J_{n-1}, J_{n}\right),\left(J_{n}, J_{1}\right)$.

Next, every two nonadjacent boxes intersect, hence their projections on both axes intersect, too. Then the claim below shows that at most 3 pairs among $\left(I_{1}, I_{2}\right), \ldots,\left(I_{n-1}, I_{n}\right),\left(I_{n}, I_{1}\right)$ are disjoint, and the same holds for $\left(J_{1}, J_{2}\right), \ldots,\left(J_{n-1}, J_{n}\right),\left(J_{n}, J_{1}\right)$. Consequently $n \leq 3+3=6$, as stated. Thus we are left with the claim and its justification.
Claim. Let $\Delta_{1}, \Delta_{2}, \ldots, \Delta_{n}$ be intervals on a straight line such that every two nonadjacent intervals intersect. Then $\Delta_{k}$ and $\Delta_{k+1}$ are disjoint for at most three values of $k=1, \ldots, n$.
Proof. Denote $\Delta_{k}=\left[a_{k}, b_{k}\right], 1 \leq k \leq n$. Let $\alpha=\max \left(a_{1}, \ldots, a_{n}\right)$ be the rightmost among the left endpoints of $\Delta_{1}, \ldots, \Delta_{n}$, and let $\beta=\min \left(b_{1}, \ldots, b_{n}\right)$ be the leftmost among their right endpoints. Assume that $\alpha=a_{2}$ without loss of generality.

If $\alpha \leq \beta$ then $a_{i} \leq \alpha \leq \beta \leq b_{i}$ for all $i$. Every $\Delta_{i}$ contains $\alpha$, and thus no disjoint pair $\left(\Delta_{i}, \Delta_{i+1}\right)$ exists.

If $\beta<\alpha$ then $\beta=b_{i}$ for some $i$ such that $a_{i}<b_{i}=\beta<\alpha=a_{2}<b_{2}$, hence $\Delta_{2}$ and $\Delta_{i}$ are disjoint. Now $\Delta_{2}$ intersects all remaining intervals except possibly $\Delta_{1}$ and $\Delta_{3}$, so $\Delta_{2}$ and $\Delta_{i}$ can be disjoint only if $i=1$ or $i=3$. Suppose by symmetry that $i=3$; then $\beta=b_{3}$. Since each of the intervals $\Delta_{4}, \ldots, \Delta_{n}$ intersects $\Delta_{2}$, we have $a_{i} \leq \alpha \leq b_{i}$ for $i=4, \ldots, n$. Therefore $\alpha \in \Delta_{4} \cap \ldots \cap \Delta_{n}$, in particular $\Delta_{4} \cap \ldots \cap \Delta_{n} \neq \emptyset$. Similarly, $\Delta_{5}, \ldots, \Delta_{n}, \Delta_{1}$ all intersect $\Delta_{3}$, so that $\Delta_{5} \cap \ldots \cap \Delta_{n} \cap \Delta_{1} \neq \emptyset$ as $\beta \in \Delta_{5} \cap \ldots \cap \Delta_{n} \cap \Delta_{1}$. This leaves $\left(\Delta_{1}, \Delta_{2}\right),\left(\Delta_{2}, \Delta_{3}\right)$ and $\left(\Delta_{3}, \Delta_{4}\right)$ as the only candidates for disjoint interval pairs, as desired.

Comment. The problem is a two-dimensional version of the original proposal which is included below. The extreme shortage of easy and appropriate submissions forced the Problem Selection Committee to shortlist a simplified variant. The same one-dimensional Claim is used in both versions.

Original proposal. We consider parallelepipeds in three-dimensional space, with edges parallel to the coordinate axes and of positive length. Such a parallelepiped will be called a box. Two boxes intersect if they have a common point in their interior or on their boundary.

Find the largest $n$ for which there exist $n$ boxes $B_{1}, \ldots, B_{n}$ such that $B_{i}$ and $B_{j}$ intersect if and only if $i \not \equiv j \pm 1(\bmod n)$.

The maximum number of such boxes is 9 . Suppose that boxes $B_{1}, \ldots, B_{n}$ satisfy the condition. Let the closed intervals $I_{k}, J_{k}$ and $K_{k}$ be the projections of box $B_{k}$ onto the $x$-, $y$ and $z$-axis, respectively, for $1 \leq k \leq n$. As before, $B_{i}$ and $B_{j}$ are disjoint if and only if their projections on at least one coordinate axis are disjoint.

We call again two boxes or intervals adjacent if their indices differ by 1 modulo $n$, and nonadjacent otherwise.

The adjacent boxes $B_{i}$ and $B_{i+1}$ do not intersect for each $i=1, \ldots, n$. Hence at least one of the pairs $\left(I_{i}, I_{i+1}\right),\left(J_{i}, J_{i+1}\right)$ and $\left(K_{i}, K_{i+1}\right)$ is a pair of disjoint intervals. So there are at least $n$ pairs of disjoint intervals among $\left(I_{i}, I_{i+1}\right),\left(J_{i}, J_{i+1}\right),\left(K_{i}, K_{i+1}\right), 1 \leq i \leq n$.

Next, every two nonadjacent boxes intersect, hence their projections on the three axes intersect, too. Referring to the Claim in the solution of the two-dimensional version, we cocnclude that at most 3 pairs among $\left(I_{1}, I_{2}\right), \ldots,\left(I_{n-1}, I_{n}\right),\left(I_{n}, I_{1}\right)$ are disjoint; the same holds for $\left(J_{1}, J_{2}\right), \ldots,\left(J_{n-1}, J_{n}\right),\left(J_{n}, J_{1}\right)$ and $\left(K_{1}, K_{2}\right), \ldots,\left(K_{n-1}, K_{n}\right),\left(K_{n}, K_{1}\right)$. Consequently $n \leq 3+3+3=9$, as stated.

For $n=9$, the desired system of boxes exists. Consider the intervals in the following table:

| $i$ | $I_{i}$ | $J_{i}$ | $K_{i}$ |
| :---: | :---: | :---: | :---: |
| 1 | $[1,4]$ | $[1,6]$ | $[3,6]$ |
| 2 | $[5,6]$ | $[1,6]$ | $[1,6]$ |
| 3 | $[1,2]$ | $[1,6]$ | $[1,6]$ |
| 4 | $[3,6]$ | $[1,4]$ | $[1,6]$ |
| 5 | $[1,6]$ | $[5,6]$ | $[1,6]$ |
| 6 | $[1,6]$ | $[1,2]$ | $[1,6]$ |
| 7 | $[1,6]$ | $[3,6]$ | $[1,4]$ |
| 8 | $[1,6]$ | $[1,6]$ | $[5,6]$ |
| 9 | $[1,6]$ | $[1,6]$ | $[1,2]$ |

We have $I_{1} \cap I_{2}=I_{2} \cap I_{3}=I_{3} \cap I_{4}=\emptyset, J_{4} \cap J_{5}=J_{5} \cap J_{6}=J_{6} \cap J_{7}=\emptyset$, and finally $K_{7} \cap K_{8}=K_{8} \cap K_{9}=K_{9} \cap K_{1}=\emptyset$. The intervals in each column intersect in all other cases. It follows that the boxes $B_{i}=I_{i} \times J_{i} \times K_{i}, i=1, \ldots, 9$, have the stated property.

C2. For every positive integer $n$ determine the number of permutations $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ of the set $\{1,2, \ldots, n\}$ with the following property:

$$
2\left(a_{1}+a_{2}+\cdots+a_{k}\right) \quad \text { is divisible by } k \text { for } k=1,2, \ldots, n \text {. }
$$

Solution. For each $n$ let $F_{n}$ be the number of permutations of $\{1,2, \ldots, n\}$ with the required property; call them nice. For $n=1,2,3$ every permutation is nice, so $F_{1}=1, F_{2}=2, F_{3}=6$.

Take an $n>3$ and consider any nice permutation $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ of $\{1,2, \ldots, n\}$. Then $n-1$ must be a divisor of the number

$$
\begin{aligned}
& 2\left(a_{1}+a_{2}+\cdots+a_{n-1}\right)=2\left((1+2+\cdots+n)-a_{n}\right) \\
& \quad=n(n+1)-2 a_{n}=(n+2)(n-1)+\left(2-2 a_{n}\right)
\end{aligned}
$$

So $2 a_{n}-2$ must be divisible by $n-1$, hence equal to 0 or $n-1$ or $2 n-2$. This means that

$$
a_{n}=1 \quad \text { or } \quad a_{n}=\frac{n+1}{2} \quad \text { or } \quad a_{n}=n
$$

Suppose that $a_{n}=(n+1) / 2$. Since the permutation is nice, taking $k=n-2$ we get that $n-2$ has to be a divisor of

$$
\begin{aligned}
2\left(a_{1}+a_{2}+\right. & \left.\cdots+a_{n-2}\right)=2\left((1+2+\cdots+n)-a_{n}-a_{n-1}\right) \\
& =n(n+1)-(n+1)-2 a_{n-1}=(n+2)(n-2)+\left(3-2 a_{n-1}\right)
\end{aligned}
$$

So $2 a_{n-1}-3$ should be divisible by $n-2$, hence equal to 0 or $n-2$ or $2 n-4$. Obviously 0 and $2 n-4$ are excluded because $2 a_{n-1}-3$ is odd. The remaining possibility ( $2 a_{n-1}-3=n-2$ ) leads to $a_{n-1}=(n+1) / 2=a_{n}$, which also cannot hold. This eliminates $(n+1) / 2$ as a possible value of $a_{n}$. Consequently $a_{n}=1$ or $a_{n}=n$.

If $a_{n}=n$ then $\left(a_{1}, a_{2}, \ldots, a_{n-1}\right)$ is a nice permutation of $\{1,2, \ldots, n-1\}$. There are $F_{n-1}$ such permutations. Attaching $n$ to any one of them at the end creates a nice permutation of $\{1,2, \ldots, n\}$.

If $a_{n}=1$ then $\left(a_{1}-1, a_{2}-1, \ldots, a_{n-1}-1\right)$ is a permutation of $\{1,2, \ldots, n-1\}$. It is also nice because the number

$$
2\left(\left(a_{1}-1\right)+\cdots+\left(a_{k}-1\right)\right)=2\left(a_{1}+\cdots+a_{k}\right)-2 k
$$

is divisible by $k$, for any $k \leq n-1$. And again, any one of the $F_{n-1}$ nice permutations $\left(b_{1}, b_{2}, \ldots, b_{n-1}\right)$ of $\{1,2, \ldots, n-1\}$ gives rise to a nice permutation of $\{1,2, \ldots, n\}$ whose last term is 1 , namely $\left(b_{1}+1, b_{2}+1, \ldots, b_{n-1}+1,1\right)$.

The bijective correspondences established in both cases show that there are $F_{n-1}$ nice permutations of $\{1,2, \ldots, n\}$ with the last term 1 and also $F_{n-1}$ nice permutations of $\{1,2, \ldots, n\}$ with the last term $n$. Hence follows the recurrence $F_{n}=2 F_{n-1}$. With the base value $F_{3}=6$ this gives the outcome formula $F_{n}=3 \cdot 2^{n-2}$ for $n \geq 3$.

C3. In the coordinate plane consider the set $S$ of all points with integer coordinates. For a positive integer $k$, two distinct points $A, B \in S$ will be called $k$-friends if there is a point $C \in S$ such that the area of the triangle $A B C$ is equal to $k$. A set $T \subset S$ will be called a $k$-clique if every two points in $T$ are $k$-friends. Find the least positive integer $k$ for which there exists a $k$-clique with more than 200 elements.

Solution. To begin, let us describe those points $B \in S$ which are $k$-friends of the point $(0,0)$. By definition, $B=(u, v)$ satisfies this condition if and only if there is a point $C=(x, y) \in S$ such that $\frac{1}{2}|u y-v x|=k$. (This is a well-known formula expressing the area of triangle $A B C$ when $A$ is the origin.)

To say that there exist integers $x, y$ for which $|u y-v x|=2 k$, is equivalent to saying that the greatest common divisor of $u$ and $v$ is also a divisor of $2 k$. Summing up, a point $B=(u, v) \in S$ is a $k$-friend of $(0,0)$ if and only if $\operatorname{gcd}(u, v)$ divides $2 k$.

Translation by a vector with integer coordinates does not affect $k$-friendship; if two points are $k$-friends, so are their translates. It follows that two points $A, B \in S, A=(s, t), B=(u, v)$, are $k$-friends if and only if the point $(u-s, v-t)$ is a $k$-friend of $(0,0)$; i.e., if $\operatorname{gcd}(u-s, v-t) \mid 2 k$.

Let $n$ be a positive integer which does not divide $2 k$. We claim that a $k$-clique cannot have more than $n^{2}$ elements.

Indeed, all points $(x, y) \in S$ can be divided into $n^{2}$ classes determined by the remainders that $x$ and $y$ leave in division by $n$. If a set $T$ has more than $n^{2}$ elements, some two points $A, B \in T, A=(s, t), B=(u, v)$, necessarily fall into the same class. This means that $n \mid u-s$ and $n \mid v-t$. Hence $n \mid d$ where $d=\operatorname{gcd}(u-s, v-t)$. And since $n$ does not divide $2 k$, also $d$ does not divide $2 k$. Thus $A$ and $B$ are not $k$-friends and the set $T$ is not a $k$-clique.

Now let $M(k)$ be the least positive integer which does not divide $2 k$. Write $M(k)=m$ for the moment and consider the set $T$ of all points $(x, y)$ with $0 \leq x, y<m$. There are $m^{2}$ of them. If $A=(s, t), B=(u, v)$ are two distinct points in $T$ then both differences $|u-s|,|v-t|$ are integers less than $m$ and at least one of them is positive. By the definition of $m$, every positive integer less than $m$ divides $2 k$. Therefore $u-s$ (if nonzero) divides $2 k$, and the same is true of $v-t$. So $2 k$ is divisible by $\operatorname{gcd}(u-s, v-t)$, meaning that $A, B$ are $k$-friends. Thus $T$ is a $k$-clique.

It follows that the maximum size of a $k$-clique is $M(k)^{2}$, with $M(k)$ defined as above. We are looking for the minimum $k$ such that $M(k)^{2}>200$.

By the definition of $M(k), 2 k$ is divisible by the numbers $1,2, \ldots, M(k)-1$, but not by $M(k)$ itself. If $M(k)^{2}>200$ then $M(k) \geq 15$. Trying to hit $M(k)=15$ we get a contradiction immediately ( $2 k$ would have to be divisible by 3 and 5 , but not by 15 ).

So let us try $M(k)=16$. Then $2 k$ is divisible by the numbers $1,2, \ldots, 15$, hence also by their least common multiple $L$, but not by 16 . And since $L$ is not a multiple of 16 , we infer that $k=L / 2$ is the least $k$ with $M(k)=16$.

Finally, observe that if $M(k) \geq 17$ then $2 k$ must be divisible by the least common multiple of $1,2, \ldots, 16$, which is equal to $2 L$. Then $2 k \geq 2 L$, yielding $k>L / 2$.

In conclusion, the least $k$ with the required property is equal to $L / 2=180180$.
$\mathbf{C 4}$. Let $n$ and $k$ be fixed positive integers of the same parity, $k \geq n$. We are given $2 n$ lamps numbered 1 through $2 n$; each of them can be on or off. At the beginning all lamps are off. We consider sequences of $k$ steps. At each step one of the lamps is switched (from off to on or from on to off).

Let $N$ be the number of $k$-step sequences ending in the state: lamps $1, \ldots, n$ on, lamps $n+1, \ldots, 2 n$ off.

Let $M$ be the number of $k$-step sequences leading to the same state and not touching lamps $n+1, \ldots, 2 n$ at all.

Find the ratio $N / M$.
Solution. A sequence of $k$ switches ending in the state as described in the problem statement (lamps $1, \ldots, n$ on, lamps $n+1, \ldots, 2 n$ off) will be called an admissible process. If, moreover, the process does not touch the lamps $n+1, \ldots, 2 n$, it will be called restricted. So there are $N$ admissible processes, among which $M$ are restricted.

In every admissible process, restricted or not, each one of the lamps $1, \ldots, n$ goes from off to on, so it is switched an odd number of times; and each one of the lamps $n+1, \ldots, 2 n$ goes from off to off, so it is switched an even number of times.

Notice that $M>0$; i.e., restricted admissible processes do exist (it suffices to switch each one of the lamps $1, \ldots, n$ just once and then choose one of them and switch it $k-n$ times, which by hypothesis is an even number).

Consider any restricted admissible process $\mathbf{p}$. Take any lamp $\ell, 1 \leq \ell \leq n$, and suppose that it was switched $k_{\ell}$ times. As noticed, $k_{\ell}$ must be odd. Select arbitrarily an even number of these $k_{\ell}$ switches and replace each of them by the switch of lamp $n+\ell$. This can be done in $2^{k_{\ell}-1}$ ways (because a $k_{\ell}$-element set has $2^{k_{\ell}-1}$ subsets of even cardinality). Notice that $k_{1}+\cdots+k_{n}=k$.

These actions are independent, in the sense that the action involving lamp $\ell$ does not affect the action involving any other lamp. So there are $2^{k_{1}-1} \cdot 2^{k_{2}-1} \cdots 2^{k_{n}-1}=2^{k-n}$ ways of combining these actions. In any of these combinations, each one of the lamps $n+1, \ldots, 2 n$ gets switched an even number of times and each one of the lamps $1, \ldots, n$ remains switched an odd number of times, so the final state is the same as that resulting from the original process $\mathbf{p}$.

This shows that every restricted admissible process $\mathbf{p}$ can be modified in $2^{k-n}$ ways, giving rise to $2^{k-n}$ distinct admissible processes (with all lamps allowed).

Now we show that every admissible process $\mathbf{q}$ can be achieved in that way. Indeed, it is enough to replace every switch of a lamp with a label $\ell>n$ that occurs in $\mathbf{q}$ by the switch of the corresponding lamp $\ell-n$; in the resulting process $\mathbf{p}$ the lamps $n+1, \ldots, 2 n$ are not involved.

Switches of each lamp with a label $\ell>n$ had occurred in $\mathbf{q}$ an even number of times. So the performed replacements have affected each lamp with a label $\ell \leq n$ also an even number of times; hence in the overall effect the final state of each lamp has remained the same. This means that the resulting process $\mathbf{p}$ is admissible - and clearly restricted, as the lamps $n+1, \ldots, 2 n$ are not involved in it any more.

If we now take process $\mathbf{p}$ and reverse all these replacements, then we obtain process $\mathbf{q}$. These reversed replacements are nothing else than the modifications described in the foregoing paragraphs.

Thus there is a one - to $-\left(2^{k-n}\right)$ correspondence between the $M$ restricted admissible processes and the total of $N$ admissible processes. Therefore $N / M=2^{k-n}$.

C5. Let $S=\left\{x_{1}, x_{2}, \ldots, x_{k+\ell}\right\}$ be a $(k+\ell)$-element set of real numbers contained in the interval $[0,1] ; k$ and $\ell$ are positive integers. A $k$-element subset $A \subset S$ is called nice if

$$
\left|\frac{1}{k} \sum_{x_{i} \in A} x_{i}-\frac{1}{\ell} \sum_{x_{j} \in S \backslash A} x_{j}\right| \leq \frac{k+\ell}{2 k \ell} .
$$

Prove that the number of nice subsets is at least $\frac{2}{k+\ell}\binom{k+\ell}{k}$.
Solution. For a $k$-element subset $A \subset S$, let $f(A)=\frac{1}{k} \sum_{x_{i} \in A} x_{i}-\frac{1}{\ell} \sum_{x_{j} \in S \backslash A} x_{j}$. Denote $\frac{k+\ell}{2 k \ell}=d$. By definition a subset $A$ is nice if $|f(A)| \leq d$.

To each permutation $\left(y_{1}, y_{2}, \ldots, y_{k+\ell}\right)$ of the set $S=\left\{x_{1}, x_{2}, \ldots, x_{k+\ell}\right\}$ we assign $k+\ell$ subsets of $S$ with $k$ elements each, namely $A_{i}=\left\{y_{i}, y_{i+1}, \ldots, y_{i+k-1}\right\}, i=1,2, \ldots, k+\ell$. Indices are taken modulo $k+\ell$ here and henceforth. In other words, if $y_{1}, y_{2}, \ldots, y_{k+\ell}$ are arranged around a circle in this order, the sets in question are all possible blocks of $k$ consecutive elements.
Claim. At least two nice sets are assigned to every permutation of $S$.
Proof. Adjacent sets $A_{i}$ and $A_{i+1}$ differ only by the elements $y_{i}$ and $y_{i+k}, i=1, \ldots, k+\ell$. By the definition of $f$, and because $y_{i}, y_{i+k} \in[0,1]$,

$$
\left|f\left(A_{i+1}\right)-f\left(A_{i}\right)\right|=\left|\left(\frac{1}{k}+\frac{1}{\ell}\right)\left(y_{i+k}-y_{i}\right)\right| \leq \frac{1}{k}+\frac{1}{\ell}=2 d
$$

Each element $y_{i} \in S$ belongs to exactly $k$ of the sets $A_{1}, \ldots, A_{k+\ell}$. Hence in $k$ of the expressions $f\left(A_{1}\right), \ldots, f\left(A_{k+\ell}\right)$ the coefficient of $y_{i}$ is $1 / k$; in the remaining $\ell$ expressions, its coefficient is $-1 / \ell$. So the contribution of $y_{i}$ to the sum of all $f\left(A_{i}\right)$ equals $k \cdot 1 / k-\ell \cdot 1 / \ell=0$. Since this holds for all $i$, it follows that $f\left(A_{1}\right)+\cdots+f\left(A_{k+\ell}\right)=0$.

If $f\left(A_{p}\right)=\min f\left(A_{i}\right), f\left(A_{q}\right)=\max f\left(A_{i}\right)$, we obtain in particular $f\left(A_{p}\right) \leq 0, f\left(A_{q}\right) \geq 0$. Let $p<q$ (the case $p>q$ is analogous; and the claim is true for $p=q$ as $f\left(A_{i}\right)=0$ for all $i$ ).

We are ready to prove that at least two of the sets $A_{1}, \ldots, A_{k+\ell}$ are nice. The interval $[-d, d]$ has length $2 d$, and we saw that adjacent numbers in the circular arrangement $f\left(A_{1}\right), \ldots, f\left(A_{k+\ell}\right)$ differ by at most $2 d$. Suppose that $f\left(A_{p}\right)<-d$ and $f\left(A_{q}\right)>d$. Then one of the numbers $f\left(A_{p+1}\right), \ldots, f\left(A_{q-1}\right)$ lies in $[-d, d]$, and also one of the numbers $f\left(A_{q+1}\right), \ldots, f\left(A_{p-1}\right)$ lies there. Consequently, one of the sets $A_{p+1}, \ldots, A_{q-1}$ is nice, as well as one of the sets $A_{q+1}, \ldots, A_{p-1}$. If $-d \leq f\left(A_{p}\right)$ and $f\left(A_{q}\right) \leq d$ then $A_{p}$ and $A_{q}$ are nice.

Let now $f\left(A_{p}\right)<-d$ and $f\left(A_{q}\right) \leq d$. Then $f\left(A_{p}\right)+f\left(A_{q}\right)<0$, and since $\sum f\left(A_{i}\right)=0$, there is an $r \neq q$ such that $f\left(A_{r}\right)>0$. We have $0<f\left(A_{r}\right) \leq f\left(A_{q}\right) \leq d$, so the sets $f\left(A_{r}\right)$ and $f\left(A_{q}\right)$ are nice. The only case remaining, $-d \leq f\left(A_{p}\right)$ and $d<f\left(A_{q}\right)$, is analogous.

Apply the claim to each of the $(k+\ell)$ ! permutations of $S=\left\{x_{1}, x_{2}, \ldots, x_{k+\ell}\right\}$. This gives at least $2(k+\ell)$ ! nice sets, counted with repetitions: each nice set is counted as many times as there are permutations to which it is assigned.

On the other hand, each $k$-element set $A \subset S$ is assigned to exactly $(k+\ell) k!\ell!$ permutations. Indeed, such a permutation $\left(y_{1}, y_{2}, \ldots, y_{k+\ell}\right)$ is determined by three independent choices: an in$\operatorname{dex} i \in\{1,2, \ldots, k+\ell\}$ such that $A=\left\{y_{i}, y_{i+1}, \ldots, y_{i+k-1}\right\}$, a permutation $\left(y_{i}, y_{i+1}, \ldots, y_{i+k-1}\right)$ of the set $A$, and a permutation $\left(y_{i+k}, y_{i+k+1}, \ldots, y_{i-1}\right)$ of the set $S \backslash A$.

In summary, there are at least $\frac{2(k+\ell)!}{(k+\ell) k!\ell!}=\frac{2}{k+\ell}\binom{k+\ell}{k}$ nice sets.

C6. For $n \geq 2$, let $S_{1}, S_{2}, \ldots, S_{2^{n}}$ be $2^{n}$ subsets of $A=\left\{1,2,3, \ldots, 2^{n+1}\right\}$ that satisfy the following property: There do not exist indices $a$ and $b$ with $a<b$ and elements $x, y, z \in A$ with $x<y<z$ such that $y, z \in S_{a}$ and $x, z \in S_{b}$. Prove that at least one of the sets $S_{1}, S_{2}, \ldots, S_{2^{n}}$ contains no more than $4 n$ elements.

Solution 1. We prove that there exists a set $S_{a}$ with at most $3 n+1$ elements.
Given a $k \in\{1, \ldots, n\}$, we say that an element $z \in A$ is $k$-good to a set $S_{a}$ if $z \in S_{a}$ and $S_{a}$ contains two other elements $x$ and $y$ with $x<y<z$ such that $z-y<2^{k}$ and $z-x \geq 2^{k}$. Also, $z \in A$ will be called good to $S_{a}$ if $z$ is $k$-good to $S_{a}$ for some $k=1, \ldots, n$.

We claim that each $z \in A$ can be $k$-good to at most one set $S_{a}$. Indeed, suppose on the contrary that $z$ is $k$-good simultaneously to $S_{a}$ and $S_{b}$, with $a<b$. Then there exist $y_{a} \in S_{a}$, $y_{a}<z$, and $x_{b} \in S_{b}, x_{b}<z$, such that $z-y_{a}<2^{k}$ and $z-x_{b} \geq 2^{k}$. On the other hand, since $z \in S_{a} \cap S_{b}$, by the condition of the problem there is no element of $S_{a}$ strictly between $x_{b}$ and $z$. Hence $y_{a} \leq x_{b}$, implying $z-y_{a} \geq z-x_{b}$. However this contradicts $z-y_{a}<2^{k}$ and $z-x_{b} \geq 2^{k}$. The claim follows.

As a consequence, a fixed $z \in A$ can be good to at most $n$ of the given sets (no more than one of them for each $k=1, \ldots, n$ ).

Furthermore, let $u_{1}<u_{2}<\cdots<u_{m}<\cdots<u_{p}$ be all elements of a fixed set $S_{a}$ that are not good to $S_{a}$. We prove that $u_{m}-u_{1}>2\left(u_{m-1}-u_{1}\right)$ for all $m \geq 3$.

Indeed, assume that $u_{m}-u_{1} \leq 2\left(u_{m-1}-u_{1}\right)$ holds for some $m \geq 3$. This inequality can be written as $2\left(u_{m}-u_{m-1}\right) \leq u_{m}-u_{1}$. Take the unique $k$ such that $2^{k} \leq u_{m}-u_{1}<2^{k+1}$. Then $2\left(u_{m}-u_{m-1}\right) \leq u_{m}-u_{1}<2^{k+1}$ yields $u_{m}-u_{m-1}<2^{k}$. However the elements $z=u_{m}, x=u_{1}$, $y=u_{m-1}$ of $S_{a}$ then satisfy $z-y<2^{k}$ and $z-x \geq 2^{k}$, so that $z=u_{m}$ is $k$-good to $S_{a}$.

Thus each term of the sequence $u_{2}-u_{1}, u_{3}-u_{1}, \ldots, u_{p}-u_{1}$ is more than twice the previous one. Hence $u_{p}-u_{1}>2^{p-1}\left(u_{2}-u_{1}\right) \geq 2^{p-1}$. But $u_{p} \in\left\{1,2,3, \ldots, 2^{n+1}\right\}$, so that $u_{p} \leq 2^{n+1}$. This yields $p-1 \leq n$, i. e. $p \leq n+1$.

In other words, each set $S_{a}$ contains at most $n+1$ elements that are not good to it.
To summarize the conclusions, mark with red all elements in the sets $S_{a}$ that are good to the respective set, and with blue the ones that are not good. Then the total number of red elements, counting multiplicities, is at most $n \cdot 2^{n+1}$ (each $z \in A$ can be marked red in at most $n$ sets). The total number of blue elements is at most $(n+1) 2^{n}$ (each set $S_{a}$ contains at most $n+1$ blue elements). Therefore the sum of cardinalities of $S_{1}, S_{2}, \ldots, S_{2^{n}}$ does not exceed $(3 n+1) 2^{n}$. By averaging, the smallest set has at most $3 n+1$ elements.

Solution 2. We show that one of the sets $S_{a}$ has at most $2 n+1$ elements. In the sequel $|\cdot|$ denotes the cardinality of a (finite) set.
Claim. For $n \geq 2$, suppose that $k$ subsets $S_{1}, \ldots, S_{k}$ of $\left\{1,2, \ldots, 2^{n}\right\}$ (not necessarily different) satisfy the condition of the problem. Then

$$
\sum_{i=1}^{k}\left(\left|S_{i}\right|-n\right) \leq(2 n-1) 2^{n-2}
$$

Proof. Observe that if the sets $S_{i}(1 \leq i \leq k)$ satisfy the condition then so do their arbitrary subsets $T_{i}(1 \leq i \leq k)$. The condition also holds for the sets $t+S_{i}=\left\{t+x \mid x \in S_{i}\right\}$ where $t$ is arbitrary.

Note also that a set may occur more than once among $S_{1}, \ldots, S_{k}$ only if its cardinality is less than 3, in which case its contribution to the sum $\sum_{i=1}^{k}\left(\left|S_{i}\right|-n\right)$ is nonpositive (as $n \geq 2$ ).

The proof is by induction on $n$. In the base case $n=2$ we have subsets $S_{i}$ of $\{1,2,3,4\}$. Only the ones of cardinality 3 and 4 need to be considered by the remark above; each one of
them occurs at most once among $S_{1}, \ldots, S_{k}$. If $S_{i}=\{1,2,3,4\}$ for some $i$ then no $S_{j}$ is a 3 -element subset in view of the condition, hence $\sum_{i=1}^{k}\left(\left|S_{i}\right|-2\right) \leq 2$. By the condition again, it is impossible that $S_{i}=\{1,3,4\}$ and $S_{j}=\{2,3,4\}$ for some $i, j$. So if $\left|S_{i}\right| \leq 3$ for all $i$ then at most 3 summands $\left|S_{i}\right|-2$ are positive, corresponding to 3 -element subsets. This implies $\sum_{i=1}^{k}\left(\left|S_{i}\right|-2\right) \leq 3$, therefore the conclusion is true for $n=2$.

Suppose that the claim holds for some $n \geq 2$, and let the sets $S_{1}, \ldots, S_{k} \subseteq\left\{1,2, \ldots, 2^{n+1}\right\}$ satisfy the given property. Denote $U_{i}=S_{i} \cap\left\{1,2, \ldots, 2^{n}\right\}, V_{i}=S_{i} \cap\left\{2^{n}+1, \ldots, 2^{n+1}\right\}$. Let

$$
I=\left\{i\left|1 \leq i \leq k,\left|U_{i}\right| \neq 0\right\}, \quad J=\{1, \ldots, k\} \backslash I\right.
$$

The sets $S_{j}$ with $j \in J$ are all contained in $\left\{2^{n}+1, \ldots, 2^{n+1}\right\}$, so the induction hypothesis applies to their translates $-2^{n}+S_{j}$ which have the same cardinalities. Consequently, this gives $\sum_{j \in J}\left(\left|S_{j}\right|-n\right) \leq(2 n-1) 2^{n-2}$, so that

$$
\begin{equation*}
\sum_{j \in J}\left(\left|S_{j}\right|-(n+1)\right) \leq \sum_{j \in J}\left(\left|S_{j}\right|-n\right) \leq(2 n-1) 2^{n-2} \tag{1}
\end{equation*}
$$

For $i \in I$, denote by $v_{i}$ the least element of $V_{i}$. Observe that if $V_{a}$ and $V_{b}$ intersect, with $a<b$, $a, b \in I$, then $v_{a}$ is their unique common element. Indeed, let $z \in V_{a} \cap V_{b} \subseteq S_{a} \cap S_{b}$ and let $m$ be the least element of $S_{b}$. Since $b \in I$, we have $m \leq 2^{n}$. By the condition, there is no element of $S_{a}$ strictly between $m \leq 2^{n}$ and $z>2^{n}$, which implies $z=v_{a}$.

It follows that if the element $v_{i}$ is removed from each $V_{i}$, a family of pairwise disjoint sets $W_{i}=V_{i} \backslash\left\{v_{i}\right\}$ is obtained, $i \in I$ (we assume $W_{i}=\emptyset$ if $V_{i}=\emptyset$ ). As $W_{i} \subseteq\left\{2^{n}+1, \ldots, 2^{n+1}\right\}$ for all $i$, we infer that $\sum_{i \in I}\left|W_{i}\right| \leq 2^{n}$. Therefore $\sum_{i \in I}\left(\left|V_{i}\right|-1\right) \leq \sum_{i \in I}\left|W_{i}\right| \leq 2^{n}$.

On the other hand, the induction hypothesis applies directly to the sets $U_{i}, i \in I$, so that $\sum_{i \in \mathcal{I}}\left(\left|U_{i}\right|-n\right) \leq(2 n-1) 2^{n-2}$. In summary,

$$
\begin{equation*}
\sum_{i \in I}\left(\left|S_{i}\right|-(n+1)\right)=\sum_{i \in I}\left(\left|U_{i}\right|-n\right)+\sum_{i \in I}\left(\left|V_{i}\right|-1\right) \leq(2 n-1) 2^{n-2}+2^{n} \tag{2}
\end{equation*}
$$

The estimates (1) and (2) are sufficient to complete the inductive step:

$$
\begin{aligned}
\sum_{i=1}^{k}\left(\left|S_{i}\right|-(n+1)\right) & =\sum_{i \in I}\left(\left|S_{i}\right|-(n+1)\right)+\sum_{j \in J}\left(\left|S_{j}\right|-(n+1)\right) \\
& \leq(2 n-1) 2^{n-2}+2^{n}+(2 n-1) 2^{n-2}=(2 n+1) 2^{n-1}
\end{aligned}
$$

Returning to the problem, consider $k=2^{n}$ subsets $S_{1}, S_{2}, \ldots, S_{2^{n}}$ of $\left\{1,2,3, \ldots, 2^{n+1}\right\}$. If they satisfy the given condition, the claim implies $\sum_{i=1}^{2^{n}}\left(\left|S_{i}\right|-(n+1)\right) \leq(2 n+1) 2^{n-1}$. By averaging again, we see that the smallest set has at most $2 n+1$ elements.

Comment. It can happen that each set $S_{i}$ has cardinality at least $n+1$. Here is an example by the proposer.

For $i=1, \ldots, 2^{n}$, let $S_{i}=\left\{i+2^{k} \mid 0 \leq k \leq n\right\}$. Then $\left|S_{i}\right|=n+1$ for all $i$. Suppose that there exist $a<b$ and $x<y<z$ such that $y, z \in S_{a}$ and $x, z \in S_{b}$. Hence $z=a+2^{k}=b+2^{l}$ for some $k>l$. Since $y \in S_{a}$ and $y<z$, we have $y \leq a+2^{k-1}$. So the element $x \in S_{b}$ satisfies

$$
x<y \leq a+2^{k-1}=z-2^{k-1} \leq z-2^{l}=b .
$$

However the least element of $S_{b}$ is $b+1$, a contradiction.

## Geometry

G1. In an acute-angled triangle $A B C$, point $H$ is the orthocentre and $A_{0}, B_{0}, C_{0}$ are the midpoints of the sides $B C, C A, A B$, respectively. Consider three circles passing through $H: \quad \omega_{a}$ around $A_{0}, \omega_{b}$ around $B_{0}$ and $\omega_{c}$ around $C_{0}$. The circle $\omega_{a}$ intersects the line $B C$ at $A_{1}$ and $A_{2} ; \omega_{b}$ intersects $C A$ at $B_{1}$ and $B_{2} ; \omega_{c}$ intersects $A B$ at $C_{1}$ and $C_{2}$. Show that the points $A_{1}, A_{2}, B_{1}, B_{2}, C_{1}, C_{2}$ lie on a circle.

Solution 1. The perpendicular bisectors of the segments $A_{1} A_{2}, B_{1} B_{2}, C_{1} C_{2}$ are also the perpendicular bisectors of $B C, C A, A B$. So they meet at $O$, the circumcentre of $A B C$. Thus $O$ is the only point that can possibly be the centre of the desired circle.

From the right triangle $O A_{0} A_{1}$ we get

$$
\begin{equation*}
O A_{1}^{2}=O A_{0}^{2}+A_{0} A_{1}^{2}=O A_{0}^{2}+A_{0} H^{2} . \tag{1}
\end{equation*}
$$

Let $K$ be the midpoint of $A H$ and let $L$ be the midpoint of $C H$. Since $A_{0}$ and $B_{0}$ are the midpoints of $B C$ and $C A$, we see that $A_{0} L \| B H$ and $B_{0} L \| A H$. Thus the segments $A_{0} L$ and $B_{0} L$ are perpendicular to $A C$ and $B C$, hence parallel to $O B_{0}$ and $O A_{0}$, respectively. Consequently $O A_{0} L B_{0}$ is a parallelogram, so that $O A_{0}$ and $B_{0} L$ are equal and parallel. Also, the midline $B_{0} L$ of triangle $A H C$ is equal and parallel to $A K$ and $K H$.

It follows that $A K A_{0} O$ and $H A_{0} O K$ are parallelograms. The first one gives $A_{0} K=O A=R$, where $R$ is the circumradius of $A B C$. From the second one we obtain

$$
\begin{equation*}
2\left(O A_{0}^{2}+A_{0} H^{2}\right)=O H^{2}+A_{0} K^{2}=O H^{2}+R^{2} \tag{2}
\end{equation*}
$$

(In a parallelogram, the sum of squares of the diagonals equals the sum of squares of the sides).
From (1) and (2) we get $O A_{1}^{2}=\left(O H^{2}+R^{2}\right) / 2$. By symmetry, the same holds for the distances $O A_{2}, O B_{1}, O B_{2}, O C_{1}$ and $O C_{2}$. Thus $A_{1}, A_{2}, B_{1}, B_{2}, C_{1}, C_{2}$ all lie on a circle with centre at $O$ and radius $\left(O H^{2}+R^{2}\right) / 2$.


Solution 2. We are going to show again that the circumcentre $O$ is equidistant from the six points in question.

Let $A^{\prime}$ be the second intersection point of $\omega_{b}$ and $\omega_{c}$. The line $B_{0} C_{0}$, which is the line of centers of circles $\omega_{b}$ and $\omega_{c}$, is a midline in triangle $A B C$, parallel to $B C$ and perpendicular to the altitude $A H$. The points $A^{\prime}$ and $H$ are symmetric with respect to the line of centers. Therefore $A^{\prime}$ lies on the line $A H$.

From the two circles $\omega_{b}$ and $\omega_{c}$ we obtain $A C_{1} \cdot A C_{2}=A A^{\prime} \cdot A H=A B_{1} \cdot A B_{2}$. So the quadrilateral $B_{1} B_{2} C_{1} C_{2}$ is cyclic. The perpendicular bisectors of the sides $B_{1} B_{2}$ and $C_{1} C_{2}$ meet at $O$. Hence $O$ is the circumcentre of $B_{1} B_{2} C_{1} C_{2}$ and so $O B_{1}=O B_{2}=O C_{1}=O C_{2}$.

Analogous arguments yield $O A_{1}=O A_{2}=O B_{1}=O B_{2}$ and $O A_{1}=O A_{2}=O C_{1}=O C_{2}$. Thus $A_{1}, A_{2}, B_{1}, B_{2}, C_{1}, C_{2}$ lie on a circle centred at $O$.


Comment. The problem can be solved without much difficulty in many ways by calculation, using trigonometry, coordinate geometry or complex numbers. As an example we present a short proof using vectors.

Solution 3. Let again $O$ and $R$ be the circumcentre and circumradius. Consider the vectors

$$
\overrightarrow{O A}=\mathbf{a}, \quad \overrightarrow{O B}=\mathbf{b}, \quad \overrightarrow{O C}=\mathbf{c}, \quad \text { where } \quad \mathbf{a}^{2}=\mathbf{b}^{2}=\mathbf{c}^{2}=R^{2}
$$

It is well known that $\overrightarrow{O H}=\mathbf{a}+\mathbf{b}+\mathbf{c}$. Accordingly,

$$
\overrightarrow{A_{0} H}=\overrightarrow{O H}-\overrightarrow{O A_{0}}=(\mathbf{a}+\mathbf{b}+\mathbf{c})-\frac{\mathbf{b}+\mathbf{c}}{2}=\frac{2 \mathbf{a}+\mathbf{b}+\mathbf{c}}{2}
$$

and

$$
\begin{gathered}
O A_{1}^{2}=O A_{0}^{2}+A_{0} A_{1}^{2}=O A_{0}^{2}+A_{0} H^{2}=\left(\frac{\mathbf{b}+\mathbf{c}}{2}\right)^{2}+\left(\frac{2 \mathbf{a}+\mathbf{b}+\mathbf{c}}{2}\right)^{2} \\
=\frac{1}{4}\left(\mathbf{b}^{2}+2 \mathbf{b} \mathbf{c}+\mathbf{c}^{2}\right)+\frac{1}{4}\left(4 \mathbf{a}^{2}+4 \mathbf{a b}+4 \mathbf{a} \mathbf{c}+\mathbf{b}^{2}+2 \mathbf{b} \mathbf{c}+\mathbf{c}^{2}\right)=2 R^{2}+(\mathbf{a b}+\mathbf{a c}+\mathbf{b c})
\end{gathered}
$$

here $\mathbf{a b}, \mathbf{b c}$, etc. denote dot products of vectors. We get the same for the distances $O A_{2}, O B_{1}$, $O B_{2}, O C_{1}$ and $O C_{2}$.

G2. Given trapezoid $A B C D$ with parallel sides $A B$ and $C D$, assume that there exist points $E$ on line $B C$ outside segment $B C$, and $F$ inside segment $A D$, such that $\angle D A E=\angle C B F$. Denote by $I$ the point of intersection of $C D$ and $E F$, and by $J$ the point of intersection of $A B$ and $E F$. Let $K$ be the midpoint of segment $E F$; assume it does not lie on line $A B$.

Prove that $I$ belongs to the circumcircle of $A B K$ if and only if $K$ belongs to the circumcircle of $C D J$.

Solution. Assume that the disposition of points is as in the diagram.
Since $\angle E B F=180^{\circ}-\angle C B F=180^{\circ}-\angle E A F$ by hypothesis, the quadrilateral $A E B F$ is cyclic. Hence $A J \cdot J B=F J \cdot J E$. In view of this equality, $I$ belongs to the circumcircle of $A B K$ if and only if $I J \cdot J K=F J \cdot J E$. Expressing $I J=I F+F J, J E=F E-F J$, and $J K=\frac{1}{2} F E-F J$, we find that $I$ belongs to the circumcircle of $A B K$ if and only if

$$
F J=\frac{I F \cdot F E}{2 I F+F E}
$$

Since $A E B F$ is cyclic and $A B, C D$ are parallel, $\angle F E C=\angle F A B=180^{\circ}-\angle C D F$. Then $C D F E$ is also cyclic, yielding $I D \cdot I C=I F \cdot I E$. It follows that $K$ belongs to the circumcircle of $C D J$ if and only if $I J \cdot I K=I F \cdot I E$. Expressing $I J=I F+F J, I K=I F+\frac{1}{2} F E$, and $I E=I F+F E$, we find that $K$ is on the circumcircle of $C D J$ if and only if

$$
F J=\frac{I F \cdot F E}{2 I F+F E}
$$

The conclusion follows.


Comment. While the figure shows $B$ inside segment $C E$, it is possible that $C$ is inside segment $B E$. Consequently, $I$ would be inside segment $E F$ and $J$ outside segment $E F$. The position of point $K$ on line $E F$ with respect to points $I, J$ may also vary.

Some case may require that an angle $\varphi$ be replaced by $180^{\circ}-\varphi$, and in computing distances, a sum may need to become a difference. All these cases can be covered by the proposed solution if it is clearly stated that signed distances and angles are used.

G3. Let $A B C D$ be a convex quadrilateral and let $P$ and $Q$ be points in $A B C D$ such that $P Q D A$ and $Q P B C$ are cyclic quadrilaterals. Suppose that there exists a point $E$ on the line segment $P Q$ such that $\angle P A E=\angle Q D E$ and $\angle P B E=\angle Q C E$. Show that the quadrilateral $A B C D$ is cyclic.

Solution 1. Let $F$ be the point on the line $A D$ such that $E F \| P A$. By hypothesis, the quadrilateral $P Q D A$ is cyclic. So if $F$ lies between $A$ and $D$ then $\angle E F D=\angle P A D=180^{\circ}-\angle E Q D$; the points $F$ and $Q$ are on distinct sides of the line $D E$ and we infer that $E F D Q$ is a cyclic quadrilateral. And if $D$ lies between $A$ and $F$ then a similar argument shows that $\angle E F D=\angle E Q D$; but now the points $F$ and $Q$ lie on the same side of $D E$, so that $E D F Q$ is a cyclic quadrilateral.

In either case we obtain the equality $\angle E F Q=\angle E D Q=\angle P A E$ which implies that $F Q \| A E$. So the triangles $E F Q$ and $P A E$ are either homothetic or parallel-congruent. More specifically, triangle $E F Q$ is the image of $P A E$ under the mapping $f$ which carries the points $P, E$ respectively to $E, Q$ and is either a homothety or translation by a vector. Note that $f$ is uniquely determined by these conditions and the position of the points $P, E, Q$ alone.

Let now $G$ be the point on the line $B C$ such that $E G \| P B$. The same reasoning as above applies to points $B, C$ in place of $A, D$, implying that the triangle $E G Q$ is the image of $P B E$ under the same mapping $f$. So $f$ sends the four points $A, P, B, E$ respectively to $F, E, G, Q$.

If $P E \neq Q E$, so that $f$ is a homothety with a centre $X$, then the lines $A F, P E, B G$-i.e. the lines $A D, P Q, B C$-are concurrent at $X$. And since $P Q D A$ and $Q P B C$ are cyclic quadrilaterals, the equalities $X A \cdot X D=X P \cdot X Q=X B \cdot X C$ hold, showing that the quadrilateral $A B C D$ is cyclic.

Finally, if $P E=Q E$, so that $f$ is a translation, then $A D\|P Q\| B C$. Thus $P Q D A$ and $Q P B C$ are isosceles trapezoids. Then also $A B C D$ is an isosceles trapezoid, hence a cyclic quadrilateral.


Solution 2. Here is another way to reach the conclusion that the lines $A D, B C$ and $P Q$ are either concurrent or parallel. From the cyclic quadrilateral $P Q D A$ we get

$$
\angle P A D=180^{\circ}-\angle P Q D=\angle Q D E+\angle Q E D=\angle P A E+\angle Q E D .
$$

Hence $\angle Q E D=\angle P A D-\angle P A E=\angle E A D$. This in view of the tangent-chord theorem means that the circumcircle of triangle $E A D$ is tangent to the line $P Q$ at $E$. Analogously, the circumcircle of triangle $E B C$ is tangent to $P Q$ at $E$.

Suppose that the line $A D$ intersects $P Q$ at $X$. Since $X E$ is tangent to the circle $(E A D)$, $X E^{2}=X A \cdot X D$. Also, $X A \cdot X D=X P \cdot X Q$ because $P, Q, D, A$ lie on a circle. Therefore $X E^{2}=X P \cdot X Q$.

It is not hard to see that this equation determines the position of the point $X$ on the line $P Q$ uniquely. Thus, if $B C$ also cuts $P Q$, say at $Y$, then the analogous equation for $Y$ yields $X=Y$, meaning that the three lines indeed concur. In this case, as well as in the case where $A D\|P Q\| B C$, the concluding argument is the same as in the first solution.

It remains to eliminate the possibility that e.g. $A D$ meets $P Q$ at $X$ while $B C \| P Q$. Indeed, $Q P B C$ would then be an isosceles trapezoid and the angle equality $\angle P B E=\angle Q C E$ would force that $E$ is the midpoint of $P Q$. So the length of $X E$, which is the geometric mean of the lengths of $X P$ and $X Q$, should also be their arithmetic mean-impossible, as $X P \neq X Q$. The proof is now complete.

Comment. After reaching the conclusion that the circles ( $E D A$ ) and ( $E B C$ ) are tangent to $P Q$ one may continue as follows. Denote the circles (PQDA), (EDA), (EBC), (QPBC) by $\omega_{1}, \omega_{2}, \omega_{3}, \omega_{4}$ respectively. Let $\ell_{i j}$ be the radical axis of the pair $\left(\omega_{i}, \omega_{j}\right)$ for $i<j$. As is well-known, the lines $\ell_{12}, \ell_{13}, \ell_{23}$ concur, possibly at infinity (let this be the meaning of the word concur in this comment). So do the lines $\ell_{12}, \ell_{14}, \ell_{24}$. Note however that $\ell_{23}$ and $\ell_{14}$ both coincide with the line $P Q$. Hence the pair $\ell_{12}, P Q$ is in both triples; thus the four lines $\ell_{12}, \ell_{13}, \ell_{24}$ and $P Q$ are concurrent.

Similarly, $\ell_{13}, \ell_{14}, \ell_{34}$ concur, $\ell_{23}, \ell_{24}, \ell_{34}$ concur, and since $\ell_{14}=\ell_{23}=P Q$, the four lines $\ell_{13}, \ell_{24}, \ell_{34}$ and $P Q$ are concurrent. The lines $\ell_{13}$ and $\ell_{24}$ are present in both quadruples, therefore all the lines $\ell_{i j}$ are concurrent. Hence the result.

G4. In an acute triangle $A B C$ segments $B E$ and $C F$ are altitudes. Two circles passing through the points $A$ and $F$ are tangent to the line $B C$ at the points $P$ and $Q$ so that $B$ lies between $C$ and $Q$. Prove that the lines $P E$ and $Q F$ intersect on the circumcircle of triangle $A E F$.

Solution 1. To approach the desired result we need some information about the slopes of the lines $P E$ and $Q F$; this information is provided by formulas (1) and (2) which we derive below.

The tangents $B P$ and $B Q$ to the two circles passing through $A$ and $F$ are equal, as $B P^{2}=B A \cdot B F=B Q^{2}$. Consider the altitude $A D$ of triangle $A B C$ and its orthocentre $H$. From the cyclic quadrilaterals $C D F A$ and $C D H E$ we get $B A \cdot B F=B C \cdot B D=B E \cdot B H$. Thus $B P^{2}=B E \cdot B H$, or $B P / B H=B E / B P$, implying that the triangles $B P H$ and $B E P$ are similar. Hence

$$
\begin{equation*}
\angle B P E=\angle B H P . \tag{1}
\end{equation*}
$$

The point $P$ lies between $D$ and $C$; this follows from the equality $B P^{2}=B C \cdot B D$. In view of this equality, and because $B P=B Q$,

$$
D P \cdot D Q=(B P-B D) \cdot(B P+B D)=B P^{2}-B D^{2}=B D \cdot(B C-B D)=B D \cdot D C
$$

Also $A D \cdot D H=B D \cdot D C$, as is seen from the similar triangles $B D H$ and $A D C$. Combining these equalities we obtain $A D \cdot D H=D P \cdot D Q$. Therefore $D H / D P=D Q / D A$, showing that the triangles $H D P$ and $Q D A$ are similar. Hence $\angle H P D=\angle Q A D$, which can be rewritten as $\angle B P H=\angle B A D+\angle B A Q$. And since $B Q$ is tangent to the circumcircle of triangle $F A Q$,

$$
\begin{equation*}
\angle B Q F=\angle B A Q=\angle B P H-\angle B A D \tag{2}
\end{equation*}
$$

From (1) and (2) we deduce

$$
\begin{aligned}
\angle B P E+\angle B Q F & =(\angle B H P+\angle B P H)-\angle B A D
\end{aligned}=\left(180^{\circ}-\angle P B H\right)-\angle B A D .
$$

Thus $\angle B P E+\angle B Q F<180^{\circ}$, which means that the rays $P E$ and $Q F$ meet. Let $S$ be the point of intersection. Then $\angle P S Q=180^{\circ}-(\angle B P E+\angle B Q F)=\angle C A B=\angle E A F$.

If $S$ lies between $P$ and $E$ then $\angle P S Q=180^{\circ}-\angle E S F$; and if $E$ lies between $P$ and $S$ then $\angle P S Q=\angle E S F$. In either case the equality $\angle P S Q=\angle E A F$ which we have obtained means that $S$ lies on the circumcircle of triangle $A E F$.


Solution 2. Let $H$ be the orthocentre of triangle $A B C$ and let $\omega$ be the circle with diameter $A H$, passing through $E$ and $F$. Introduce the points of intersection of $\omega$ with the following lines emanating from $P: P A \cap \omega=\{A, U\}, P H \cap \omega=\{H, V\}, P E \cap \omega=\{E, S\}$. The altitudes of triangle $A H P$ are contained in the lines $A V, H U, B C$, meeting at its orthocentre $Q^{\prime}$.

By Pascal's theorem applied to the (tied) hexagon $A E S F H V$, the points $A E \cap F H=C$, $E S \cap H V=P$ and $S F \cap V A$ are collinear, so $F S$ passes through $Q^{\prime}$.

Denote by $\omega_{1}$ and $\omega_{2}$ the circles with diameters $B C$ and $P Q^{\prime}$, respectively. Let $D$ be the foot of the altitude from $A$ in triangle $A B C$. Suppose that $A D$ meets the circles $\omega_{1}$ and $\omega_{2}$ at the respective points $K$ and $L$.

Since $H$ is the orthocentre of $A B C$, the triangles $B D H$ and $A D C$ are similar, and so $D A \cdot D H=D B \cdot D C=D K^{2}$; the last equality holds because $B K C$ is a right triangle. Since $H$ is the orthocentre also in triangle $A Q^{\prime} P$, we analogously have $D L^{2}=D A \cdot D H$. Therefore $D K=D L$ and $K=L$.

Also, $B D \cdot B C=B A \cdot B F$, from the similar triangles $A B D, C B F$. In the right triangle $B K C$ we have $B K^{2}=B D \cdot B C$. Hence, and because $B A \cdot B F=B P^{2}=B Q^{2}$ (by the definition of $P$ and $Q$ in the problem statement), we obtain $B K=B P=B Q$. It follows that $B$ is the centre of $\omega_{2}$ and hence $Q^{\prime}=Q$. So the lines $P E$ and $Q F$ meet at the point $S$ lying on the circumcircle of triangle $A E F$.


Comment 1. If $T$ is the point defined by $P F \cap \omega=\{F, T\}$, Pascal's theorem for the hexagon $A F T E H V$ will analogously lead to the conclusion that the line $E T$ goes through $Q^{\prime}$. In other words, the lines $P F$ and $Q E$ also concur on $\omega$.

Comment 2. As is known from algebraic geometry, the points of the circle $\omega$ form a commutative groups with the operation defined as follows. Choose any point $0 \in \omega$ (to be the neutral element of the group) and a line $\ell$ exterior to the circle. For $X, Y \in \omega$, draw the line from the point $X Y \cap \ell$ through 0 to its second intersection with $\omega$ and define this point to be $X+Y$.

In our solution we have chosen $H$ to be the neutral element in this group and line $B C$ to be $\ell$. The fact that the lines $A V, H U, E T, F S$ are concurrent can be deduced from the identities $A+A=0$, $F=E+A, \quad V=U+A=S+E=T+F$.

Comment 3. The problem was submitted in the following equivalent formulation:
Let $B E$ and $C F$ be altitudes of an acute triangle $A B C$. We choose $P$ on the side $B C$ and $Q$ on the extension of $C B$ beyond $B$ such that $B Q^{2}=B P^{2}=B F \cdot A B$. If $Q F$ and $P E$ intersect at $S$, prove that $E S A F$ is cyclic.

G5. Let $k$ and $n$ be integers with $0 \leq k \leq n-2$. Consider a set $L$ of $n$ lines in the plane such that no two of them are parallel and no three have a common point. Denote by $I$ the set of intersection points of lines in $L$. Let $O$ be a point in the plane not lying on any line of $L$.

A point $X \in I$ is colored red if the open line segment $O X$ intersects at most $k$ lines in $L$. Prove that $I$ contains at least $\frac{1}{2}(k+1)(k+2)$ red points.

Solution. There are at least $\frac{1}{2}(k+1)(k+2)$ points in the intersection set $I$ in view of the condition $n \geq k+2$.

For each point $P \in I$, define its order as the number of lines that intersect the open line segment $O P$. By definition, $P$ is red if its order is at most $k$. Note that there is always at least one point $X \in I$ of order 0 . Indeed, the lines in $L$ divide the plane into regions, bounded or not, and $O$ belongs to one of them. Clearly any corner of this region is a point of $I$ with order 0 .
Claim. Suppose that two points $P, Q \in I$ lie on the same line of $L$, and no other line of $L$ intersects the open line segment $P Q$. Then the orders of $P$ and $Q$ differ by at most 1 .
Proof. Let $P$ and $Q$ have orders $p$ and $q$, respectively, with $p \geq q$. Consider triangle $O P Q$. Now $p$ equals the number of lines in $L$ that intersect the interior of side $O P$. None of these lines intersects the interior of side $P Q$, and at most one can pass through $Q$. All remaining lines must intersect the interior of side $O Q$, implying that $q \geq p-1$. The conclusion follows.

We prove the main result by induction on $k$. The base $k=0$ is clear since there is a point of order 0 which is red. Assuming the statement true for $k-1$, we pass on to the inductive step. Select a point $P \in I$ of order 0 , and consider one of the lines $\ell \in L$ that pass through $P$. There are $n-1$ intersection points on $\ell$, one of which is $P$. Out of the remaining $n-2$ points, the $k$ closest to $P$ have orders not exceeding $k$ by the Claim. It follows that there are at least $k+1$ red points on $\ell$.

Let us now consider the situation with $\ell$ removed (together with all intersection points it contains). By hypothesis of induction, there are at least $\frac{1}{2} k(k+1)$ points of order not exceeding $k-1$ in the resulting configuration. Restoring $\ell$ back produces at most one new intersection point on each line segment joining any of these points to $O$, so their order is at most $k$ in the original configuration. The total number of points with order not exceeding $k$ is therefore at least $(k+1)+\frac{1}{2} k(k+1)=\frac{1}{2}(k+1)(k+2)$. This completes the proof.

Comment. The steps of the proof can be performed in reverse order to obtain a configuration of $n$ lines such that equality holds simultaneously for all $0 \leq k \leq n-2$. Such a set of lines is illustrated in the Figure.


G6. There is given a convex quadrilateral $A B C D$. Prove that there exists a point $P$ inside the quadrilateral such that

$$
\begin{equation*}
\angle P A B+\angle P D C=\angle P B C+\angle P A D=\angle P C D+\angle P B A=\angle P D A+\angle P C B=90^{\circ} \tag{1}
\end{equation*}
$$

if and only if the diagonals $A C$ and $B D$ are perpendicular.
Solution 1. For a point $P$ in $A B C D$ which satisfies (1), let $K, L, M, N$ be the feet of perpendiculars from $P$ to lines $A B, B C, C D, D A$, respectively. Note that $K, L, M, N$ are interior to the sides as all angles in (1) are acute. The cyclic quadrilaterals $A K P N$ and $D N P M$ give

$$
\angle P A B+\angle P D C=\angle P N K+\angle P N M=\angle K N M
$$

Analogously, $\angle P B C+\angle P A D=\angle L K N$ and $\angle P C D+\angle P B A=\angle M L K$. Hence the equalities (1) imply $\angle K N M=\angle L K N=\angle M L K=90^{\circ}$, so that $K L M N$ is a rectangle. The converse also holds true, provided that $K, L, M, N$ are interior to sides $A B, B C, C D, D A$.
(i) Suppose that there exists a point $P$ in $A B C D$ such that $K L M N$ is a rectangle. We show that $A C$ and $B D$ are parallel to the respective sides of $K L M N$.

Let $O_{A}$ and $O_{C}$ be the circumcentres of the cyclic quadrilaterals $A K P N$ and $C M P L$. Line $O_{A} O_{C}$ is the common perpendicular bisector of $L M$ and $K N$, therefore $O_{A} O_{C}$ is parallel to $K L$ and $M N$. On the other hand, $O_{A} O_{C}$ is the midline in the triangle $A C P$ that is parallel to $A C$. Therefore the diagonal $A C$ is parallel to the sides $K L$ and $M N$ of the rectangle. Likewise, $B D$ is parallel to $K N$ and $L M$. Hence $A C$ and $B D$ are perpendicular.

(ii) Suppose that $A C$ and $B D$ are perpendicular and meet at $R$. If $A B C D$ is a rhombus, $P$ can be chosen to be its centre. So assume that $A B C D$ is not a rhombus, and let $B R<D R$ without loss of generality.

Denote by $U_{A}$ and $U_{C}$ the circumcentres of the triangles $A B D$ and $C D B$, respectively. Let $A V_{A}$ and $C V_{C}$ be the diameters through $A$ and $C$ of the two circumcircles. Since $A R$ is an altitude in triangle $A D B$, lines $A C$ and $A V_{A}$ are isogonal conjugates, i. e. $\angle D A V_{A}=\angle B A C$. Now $B R<D R$ implies that ray $A U_{A}$ lies in $\angle D A C$. Similarly, ray $C U_{C}$ lies in $\angle D C A$. Both diameters $A V_{A}$ and $C V_{C}$ intersect $B D$ as the angles at $B$ and $D$ of both triangles are acute. Also $U_{A} U_{C}$ is parallel to $A C$ as it is the perpendicular bisector of $B D$. Hence $V_{A} V_{C}$ is parallel to $A C$, too. We infer that $A V_{A}$ and $C V_{C}$ intersect at a point $P$ inside triangle $A C D$, hence inside $A B C D$.

Construct points $K, L, M, N, O_{A}$ and $O_{C}$ in the same way as in the introduction. It follows from the previous paragraph that $K, L, M, N$ are interior to the respective sides. Now $O_{A} O_{C}$ is a midline in triangle $A C P$ again. Therefore lines $A C, O_{A} O_{C}$ and $U_{A} U_{C}$ are parallel.

The cyclic quadrilateral $A K P N$ yields $\angle N K P=\angle N A P$. Since $\angle N A P=\angle D A U_{A}=$ $\angle B A C$, as specified above, we obtain $\angle N K P=\angle B A C$. Because $P K$ is perpendicular to $A B$, it follows that $N K$ is perpendicular to $A C$, hence parallel to $B D$. Likewise, $L M$ is parallel to $B D$.

Consider the two homotheties with centres $A$ and $C$ which transform triangles $A B D$ and $C D B$ into triangles $A K N$ and $C M L$, respectively. The images of points $U_{A}$ and $U_{C}$ are $O_{A}$ and $O_{C}$, respectively. Since $U_{A} U_{C}$ and $O_{A} O_{C}$ are parallel to $A C$, the two ratios of homothety are the same, equal to $\lambda=A N / A D=A K / A B=A O_{A} / A U_{A}=C O_{C} / C U_{C}=C M / C D=C L / C B$. It is now straightforward that $D N / D A=D M / D C=B K / B A=B L / B C=1-\lambda$. Hence $K L$ and $M N$ are parallel to $A C$, implying that $K L M N$ is a rectangle and completing the proof.


Solution 2. For a point $P$ distinct from $A, B, C, D$, let circles $(A P D)$ and ( $B P C$ ) intersect again at $Q(Q=P$ if the circles are tangent). Next, let circles $(A Q B)$ and $(C Q D)$ intersect again at $R$. We show that if $P$ lies in $A B C D$ and satisfies (1) then $A C$ and $B D$ intersect at $R$ and are perpendicular; the converse is also true. It is convenient to use directed angles. Let $\measuredangle(U V, X Y)$ denote the angle of counterclockwise rotation that makes line $U V$ parallel to line $X Y$. Recall that four noncollinear points $U, V, X, Y$ are concyclic if and only if $\measuredangle(U X, V X)=\measuredangle(U Y, V Y)$.

The definitions of points $P, Q$ and $R$ imply

$$
\begin{aligned}
\measuredangle(A R, B R) & =\measuredangle(A Q, B Q)=\measuredangle(A Q, P Q)+\measuredangle(P Q, B Q)=\measuredangle(A D, P D)+\measuredangle(P C, B C), \\
\measuredangle(C R, D R) & =\measuredangle(C Q, D Q)=\measuredangle(C Q, P Q)+\measuredangle(P Q, D Q)=\measuredangle(C B, P B)+\measuredangle(P A, D A), \\
\measuredangle(B R, C R) & =\measuredangle(B R, R Q)+\measuredangle(R Q, C R)=\measuredangle(B A, A Q)+\measuredangle(D Q, C D) \\
& =\measuredangle(B A, A P)+\measuredangle(A P, A Q)+\measuredangle(D Q, D P)+\measuredangle(D P, C D) \\
& =\measuredangle(B A, A P)+\measuredangle(D P, C D) .
\end{aligned}
$$

Observe that the whole construction is reversible. One may start with point $R$, define $Q$ as the second intersection of circles $(A R B)$ and $(C R D)$, and then define $P$ as the second intersection of circles $(A Q D)$ and $(B Q C)$. The equalities above will still hold true.

Assume in addition that $P$ is interior to $A B C D$. Then

$$
\begin{gathered}
\measuredangle(A D, P D)=\angle P D A, \measuredangle(P C, B C)=\angle P C B, \measuredangle(C B, P B)=\angle P B C, \measuredangle(P A, D A)=\angle P A D, \\
\measuredangle(B A, A P)=\angle P A B, \measuredangle(D P, C D)=\angle P D C .
\end{gathered}
$$

(i) Suppose that $P$ lies in $A B C D$ and satisfies (1). Then $\measuredangle(A R, B R)=\angle P D A+\angle P C B=90^{\circ}$ and similarly $\measuredangle(B R, C R)=\measuredangle(C R, D R)=90^{\circ}$. It follows that $R$ is the common point of lines $A C$ and $B D$, and that these lines are perpendicular.
(ii) Suppose that $A C$ and $B D$ are perpendicular and intersect at $R$. We show that the point $P$ defined by the reverse construction (starting with $R$ and ending with $P$ ) lies in $A B C D$. This is enough to finish the solution, because then the angle equalities above will imply (1).

One can assume that $Q$, the second common point of circles $(A B R)$ and $(C D R)$, lies in $\angle A R D$. Then in fact $Q$ lies in triangle $A D R$ as angles $A Q R$ and $D Q R$ are obtuse. Hence $\angle A Q D$ is obtuse, too, so that $B$ and $C$ are outside circle $(A D Q)(\angle A B D$ and $\angle A C D$ are acute).

Now $\angle C A B+\angle C D B=\angle B Q R+\angle C Q R=\angle C Q B$ implies $\angle C A B<\angle C Q B$ and $\angle C D B<$ $\angle C Q B$. Hence $A$ and $D$ are outside circle ( $B C Q$ ). In conclusion, the second common point $P$ of circles $(A D Q)$ and $(B C Q)$ lies on their arcs $A D Q$ and $B C Q$.

We can assume that $P$ lies in $\angle C Q D$. Since

$$
\begin{gathered}
\angle Q P C+\angle Q P D=\left(180^{\circ}-\angle Q B C\right)+\left(180^{\circ}-\angle Q A D\right)= \\
=360^{\circ}-(\angle R B C+\angle Q B R)-(\angle R A D-\angle Q A R)=360^{\circ}-\angle R B C-\angle R A D>180^{\circ},
\end{gathered}
$$

point $P$ lies in triangle $C D Q$, and hence in $A B C D$. The proof is complete.


G7. Let $A B C D$ be a convex quadrilateral with $A B \neq B C$. Denote by $\omega_{1}$ and $\omega_{2}$ the incircles of triangles $A B C$ and $A D C$. Suppose that there exists a circle $\omega$ inscribed in angle $A B C$, tangent to the extensions of line segments $A D$ and $C D$. Prove that the common external tangents of $\omega_{1}$ and $\omega_{2}$ intersect on $\omega$.

Solution. The proof below is based on two known facts.
Lemma 1. Given a convex quadrilateral $A B C D$, suppose that there exists a circle which is inscribed in angle $A B C$ and tangent to the extensions of line segments $A D$ and $C D$. Then $A B+A D=C B+C D$.
Proof. The circle in question is tangent to each of the lines $A B, B C, C D, D A$, and the respective points of tangency $K, L, M, N$ are located as with circle $\omega$ in the figure. Then

$$
A B+A D=(B K-A K)+(A N-D N), \quad C B+C D=(B L-C L)+(C M-D M)
$$

Also $B K=B L, D N=D M, A K=A N, C L=C M$ by equalities of tangents. It follows that $A B+A D=C B+C D$.


For brevity, in the sequel we write "excircle $A C$ " for the excircle of a triangle with side $A C$ which is tangent to line segment $A C$ and the extensions of the other two sides.

Lemma 2. The incircle of triangle $A B C$ is tangent to its side $A C$ at $P$. Let $P P^{\prime}$ be the diameter of the incircle through $P$, and let line $B P^{\prime}$ intersect $A C$ at $Q$. Then $Q$ is the point of tangency of side $A C$ and excircle $A C$.

Proof. Let the tangent at $P^{\prime}$ to the incircle $\omega_{1}$ meet $B A$ and $B C$ at $A^{\prime}$ and $C^{\prime}$. Now $\omega_{1}$ is the excircle $A^{\prime} C^{\prime}$ of triangle $A^{\prime} B C^{\prime}$, and it touches side $A^{\prime} C^{\prime}$ at $P^{\prime}$. Since $A^{\prime} C^{\prime} \| A C$, the homothety with centre $B$ and ratio $B Q / B P^{\prime}$ takes $\omega_{1}$ to the excircle $A C$ of triangle $A B C$. Because this homothety takes $P^{\prime}$ to $Q$, the lemma follows.

Recall also that if the incircle of a triangle touches its side $A C$ at $P$, then the tangency point $Q$ of the same side and excircle $A C$ is the unique point on line segment $A C$ such that $A P=C Q$.

We pass on to the main proof. Let $\omega_{1}$ and $\omega_{2}$ touch $A C$ at $P$ and $Q$, respectively; then $A P=(A C+A B-B C) / 2, C Q=(C A+C D-A D) / 2$. Since $A B-B C=C D-A D$ by Lemma 1, we obtain $A P=C Q$. It follows that in triangle $A B C$ side $A C$ and excircle $A C$ are tangent at $Q$. Likewise, in triangle $A D C$ side $A C$ and excircle $A C$ are tangent at $P$. Note that $P \neq Q$ as $A B \neq B C$.

Let $P P^{\prime}$ and $Q Q^{\prime}$ be the diameters perpendicular to $A C$ of $\omega_{1}$ and $\omega_{2}$, respectively. Then Lemma 2 shows that points $B, P^{\prime}$ and $Q$ are collinear, and so are points $D, Q^{\prime}$ and $P$.

Consider the diameter of $\omega$ perpendicular to $A C$ and denote by $T$ its endpoint that is closer to $A C$. The homothety with centre $B$ and ratio $B T / B P^{\prime}$ takes $\omega_{1}$ to $\omega$. Hence $B, P^{\prime}$ and $T$ are collinear. Similarly, $D, Q^{\prime}$ and $T$ are collinear since the homothety with centre $D$ and ratio $-D T / D Q^{\prime}$ takes $\omega_{2}$ to $\omega$.

We infer that points $T, P^{\prime}$ and $Q$ are collinear, as well as $T, Q^{\prime}$ and $P$. Since $P P^{\prime} \| Q Q^{\prime}$, line segments $P P^{\prime}$ and $Q Q^{\prime}$ are then homothetic with centre $T$. The same holds true for circles $\omega_{1}$ and $\omega_{2}$ because they have $P P^{\prime}$ and $Q Q^{\prime}$ as diameters. Moreover, it is immediate that $T$ lies on the same side of line $P P^{\prime}$ as $Q$ and $Q^{\prime}$, hence the ratio of homothety is positive. In particular $\omega_{1}$ and $\omega_{2}$ are not congruent.

In summary, $T$ is the centre of a homothety with positive ratio that takes circle $\omega_{1}$ to circle $\omega_{2}$. This completes the solution, since the only point with the mentioned property is the intersection of the the common external tangents of $\omega_{1}$ and $\omega_{2}$.

## Number Theory

N1. Let $n$ be a positive integer and let $p$ be a prime number. Prove that if $a, b, c$ are integers (not necessarily positive) satisfying the equations

$$
a^{n}+p b=b^{n}+p c=c^{n}+p a,
$$

then $a=b=c$.
Solution 1. If two of $a, b, c$ are equal, it is immediate that all the three are equal. So we may assume that $a \neq b \neq c \neq a$. Subtracting the equations we get $a^{n}-b^{n}=-p(b-c)$ and two cyclic copies of this equation, which upon multiplication yield

$$
\begin{equation*}
\frac{a^{n}-b^{n}}{a-b} \cdot \frac{b^{n}-c^{n}}{b-c} \cdot \frac{c^{n}-a^{n}}{c-a}=-p^{3} . \tag{1}
\end{equation*}
$$

If $n$ is odd then the differences $a^{n}-b^{n}$ and $a-b$ have the same sign and the product on the left is positive, while $-p^{3}$ is negative. So $n$ must be even.

Let $d$ be the greatest common divisor of the three differences $a-b, b-c, c-a$, so that $a-b=d u, b-c=d v, c-a=d w ; \quad \operatorname{ccd}(u, v, w)=1, u+v+w=0$.

From $a^{n}-b^{n}=-p(b-c)$ we see that $(a-b) \mid p(b-c)$, i.e., $u \mid p v$; and cyclically $v|p w, w| p u$. As $\operatorname{gcd}(u, v, w)=1$ and $u+v+w=0$, at most one of $u, v, w$ can be divisible by $p$. Supposing that the prime $p$ does not divide any one of them, we get $u|v, v| w, w \mid u$, whence $|u|=|v|=|w|=1$; but this quarrels with $u+v+w=0$.

Thus $p$ must divide exactly one of these numbers. Let e.g. $p \mid u$ and write $u=p u_{1}$. Now we obtain, similarly as before, $u_{1}|v, v| w, w \mid u_{1}$ so that $\left|u_{1}\right|=|v|=|w|=1$. The equation $p u_{1}+v+w=0$ forces that the prime $p$ must be even; i.e. $p=2$. Hence $v+w=-2 u_{1}= \pm 2$, implying $v=w(= \pm 1)$ and $u=-2 v$. Consequently $a-b=-2(b-c)$.

Knowing that $n$ is even, say $n=2 k$, we rewrite the equation $a^{n}-b^{n}=-p(b-c)$ with $p=2$ in the form

$$
\left(a^{k}+b^{k}\right)\left(a^{k}-b^{k}\right)=-2(b-c)=a-b .
$$

The second factor on the left is divisible by $a-b$, so the first factor $\left(a^{k}+b^{k}\right)$ must be $\pm 1$. Then exactly one of $a$ and $b$ must be odd; yet $a-b=-2(b-c)$ is even. Contradiction ends the proof.

Solution 2. The beginning is as in the first solution. Assuming that $a, b, c$ are not all equal, hence are all distinct, we derive equation (1) with the conclusion that $n$ is even. Write $n=2 k$.

Suppose that $p$ is odd. Then the integer

$$
\frac{a^{n}-b^{n}}{a-b}=a^{n-1}+a^{n-2} b+\cdots+b^{n-1}
$$

which is a factor in (1), must be odd as well. This sum of $n=2 k$ summands is odd only if $a$ and $b$ have different parities. The same conclusion holding for $b, c$ and for $c, a$, we get that $a, b, c, a$ alternate in their parities, which is clearly impossible.

Thus $p=2$. The original system shows that $a, b, c$ must be of the same parity. So we may divide (1) by $p^{3}$, i.e. $2^{3}$, to obtain the following product of six integer factors:

$$
\begin{equation*}
\frac{a^{k}+b^{k}}{2} \cdot \frac{a^{k}-b^{k}}{a-b} \cdot \frac{b^{k}+c^{k}}{2} \cdot \frac{b^{k}-c^{k}}{b-c} \cdot \frac{c^{k}+a^{k}}{2} \cdot \frac{c^{k}-a^{k}}{c-a}=-1 \tag{2}
\end{equation*}
$$

Each one of the factors must be equal to $\pm 1$. In particular, $a^{k}+b^{k}= \pm 2$. If $k$ is even, this becomes $a^{k}+b^{k}=2$ and yields $|a|=|b|=1$, whence $a^{k}-b^{k}=0$, contradicting (2).

Let now $k$ be odd. Then the sum $a^{k}+b^{k}$, with value $\pm 2$, has $a+b$ as a factor. Since $a$ and $b$ are of the same parity, this means that $a+b= \pm 2$; and cyclically, $b+c= \pm 2, c+a= \pm 2$. In some two of these equations the signs must coincide, hence some two of $a, b, c$ are equal. This is the desired contradiction.

Comment. Having arrived at the equation (1) one is tempted to write down all possible decompositions of $-p^{3}$ (cube of a prime) into a product of three integers. This leads to cumbersome examination of many cases, some of which are unpleasant to handle. One may do that just for $p=2$, having earlier in some way eliminated odd primes from consideration.

However, the second solution shows that the condition of $p$ being a prime is far too strong. What is actually being used in that solution, is that $p$ is either a positive odd integer or $p=2$.

N2. Let $a_{1}, a_{2}, \ldots, a_{n}$ be distinct positive integers, $n \geq 3$. Prove that there exist distinct indices $i$ and $j$ such that $a_{i}+a_{j}$ does not divide any of the numbers $3 a_{1}, 3 a_{2}, \ldots, 3 a_{n}$.

Solution. Without loss of generality, let $0<a_{1}<a_{2}<\cdots<a_{n}$. One can also assume that $a_{1}, a_{2}, \ldots, a_{n}$ are coprime. Otherwise division by their greatest common divisor reduces the question to the new sequence whose terms are coprime integers.

Suppose that the claim is false. Then for each $i<n$ there exists a $j$ such that $a_{n}+a_{i}$ divides $3 a_{j}$. If $a_{n}+a_{i}$ is not divisible by 3 then $a_{n}+a_{i}$ divides $a_{j}$ which is impossible as $0<a_{j} \leq a_{n}<a_{n}+a_{i}$. Thus $a_{n}+a_{i}$ is a multiple of 3 for $i=1, \ldots, n-1$, so that $a_{1}, a_{2}, \ldots, a_{n-1}$ are all congruent (to $-a_{n}$ ) modulo 3 .

Now $a_{n}$ is not divisible by 3 or else so would be all remaining $a_{i}$ 's, meaning that $a_{1}, a_{2}, \ldots, a_{n}$ are not coprime. Hence $a_{n} \equiv r(\bmod 3)$ where $r \in\{1,2\}$, and $a_{i} \equiv 3-r(\bmod 3)$ for all $i=1, \ldots, n-1$.

Consider a sum $a_{n-1}+a_{i}$ where $1 \leq i \leq n-2$. There is at least one such sum as $n \geq 3$. Let $j$ be an index such that $a_{n-1}+a_{i}$ divides $3 a_{j}$. Observe that $a_{n-1}+a_{i}$ is not divisible by 3 since $a_{n-1}+a_{i} \equiv 2 a_{i} \not \equiv 0(\bmod 3)$. It follows that $a_{n-1}+a_{i}$ divides $a_{j}$, in particular $a_{n-1}+a_{i} \leq a_{j}$. Hence $a_{n-1}<a_{j} \leq a_{n}$, implying $j=n$. So $a_{n}$ is divisible by all sums $a_{n-1}+a_{i}, 1 \leq i \leq n-2$. In particular $a_{n-1}+a_{i} \leq a_{n}$ for $i=1, \ldots, n-2$.

Let $j$ be such that $a_{n}+a_{n-1}$ divides $3 a_{j}$. If $j \leq n-2$ then $a_{n}+a_{n-1} \leq 3 a_{j}<a_{j}+2 a_{n-1}$. This yields $a_{n}<a_{n-1}+a_{j}$; however $a_{n-1}+a_{j} \leq a_{n}$ for $j \leq n-2$. Therefore $j=n-1$ or $j=n$.

For $j=n-1$ we obtain $3 a_{n-1}=k\left(a_{n}+a_{n-1}\right)$ with $k$ an integer, and it is straightforward that $k=1\left(k \leq 0\right.$ and $k \geq 3$ contradict $0<a_{n-1}<a_{n} ; k=2$ leads to $\left.a_{n-1}=2 a_{n}>a_{n-1}\right)$. Thus $3 a_{n-1}=a_{n}+a_{n-1}$, i. e. $a_{n}=2 a_{n-1}$.

Similarly, if $j=n$ then $3 a_{n}=k\left(a_{n}+a_{n-1}\right)$ for some integer $k$, and only $k=2$ is possible. Hence $a_{n}=2 a_{n-1}$ holds true in both cases remaining, $j=n-1$ and $j=n$.

Now $a_{n}=2 a_{n-1}$ implies that the sum $a_{n-1}+a_{1}$ is strictly between $a_{n} / 2$ and $a_{n}$. But $a_{n-1}$ and $a_{1}$ are distinct as $n \geq 3$, so it follows from the above that $a_{n-1}+a_{1}$ divides $a_{n}$. This provides the desired contradiction.

N3. Let $a_{0}, a_{1}, a_{2}, \ldots$ be a sequence of positive integers such that the greatest common divisor of any two consecutive terms is greater than the preceding term; in symbols, $\operatorname{gcd}\left(a_{i}, a_{i+1}\right)>a_{i-1}$. Prove that $a_{n} \geq 2^{n}$ for all $n \geq 0$.

Solution. Since $a_{i} \geq \operatorname{gcd}\left(a_{i}, a_{i+1}\right)>a_{i-1}$, the sequence is strictly increasing. In particular $a_{0} \geq 1, a_{1} \geq 2$. For each $i \geq 1$ we also have $a_{i+1}-a_{i} \geq \operatorname{gcd}\left(a_{i}, a_{i+1}\right)>a_{i-1}$, and consequently $a_{i+1} \geq a_{i}+a_{i-1}+1$. Hence $a_{2} \geq 4$ and $a_{3} \geq 7$. The equality $a_{3}=7$ would force equalities in the previous estimates, leading to $\operatorname{gcd}\left(a_{2}, a_{3}\right)=\operatorname{gcd}(4,7)>a_{1}=2$, which is false. Thus $a_{3} \geq 8$; the result is valid for $n=0,1,2,3$. These are the base cases for a proof by induction.

Take an $n \geq 3$ and assume that $a_{i} \geq 2^{i}$ for $i=0,1, \ldots, n$. We must show that $a_{n+1} \geq 2^{n+1}$. Let $\operatorname{gcd}\left(a_{n}, a_{n+1}\right)=d$. We know that $d>a_{n-1}$. The induction claim is reached immediately in the following cases:

$$
\begin{aligned}
& \text { if } a_{n+1} \geq 4 d \text { then } a_{n+1}>4 a_{n-1} \geq 4 \cdot 2^{n-1}=2^{n+1} \\
& \text { if } a_{n} \geq 3 d \quad \text { then } a_{n+1} \geq a_{n}+d \geq 4 d>4 a_{n-1} \geq 4 \cdot 2^{n-1}=2^{n+1} ; \\
& \text { if } a_{n}=d \quad \text { then } a_{n+1} \geq a_{n}+d=2 a_{n} \geq 2 \cdot 2^{n}=2^{n+1} .
\end{aligned}
$$

The only remaining possibility is that $a_{n}=2 d$ and $a_{n+1}=3 d$, which we assume for the sequel. So $a_{n+1}=\frac{3}{2} a_{n}$.

Let now $\operatorname{gcd}\left(a_{n-1}, a_{n}\right)=d^{\prime}$; then $d^{\prime}>a_{n-2}$. Write $a_{n}=m d^{\prime}$ ( $m$ an integer). Keeping in mind that $d^{\prime} \leq a_{n-1}<d$ and $a_{n}=2 d$, we get that $m \geq 3$. Also $a_{n-1}<d=\frac{1}{2} m d^{\prime}$, $a_{n+1}=\frac{3}{2} m d^{\prime}$. Again we single out the cases which imply the induction claim immediately:

$$
\begin{aligned}
& \text { if } m \geq 6 \quad \text { then } a_{n+1}=\frac{3}{2} m d^{\prime} \geq 9 d^{\prime}>9 a_{n-2} \geq 9 \cdot 2^{n-2}>2^{n+1} ; \\
& \text { if } 3 \leq m \leq 4 \text { then } a_{n-1}<\frac{1}{2} \cdot 4 d^{\prime}, \text { and hence } a_{n-1}=d^{\prime} \\
& \qquad a_{n+1}=\frac{3}{2} m a_{n-1} \geq \frac{3}{2} \cdot 3 a_{n-1} \geq \frac{9}{2} \cdot 2^{n-1}>2^{n+1}
\end{aligned}
$$

So we are left with the case $m=5$, which means that $a_{n}=5 d^{\prime}, a_{n+1}=\frac{15}{2} d^{\prime}, a_{n-1}<d=\frac{5}{2} d^{\prime}$. The last relation implies that $a_{n-1}$ is either $d^{\prime}$ or $2 d^{\prime}$. Anyway, $a_{n-1} \mid 2 d^{\prime}$.

The same pattern repeats once more. We denote $\operatorname{gcd}\left(a_{n-2}, a_{n-1}\right)=d^{\prime \prime}$; then $d^{\prime \prime}>a_{n-3}$. Because $d^{\prime \prime}$ is a divisor of $a_{n-1}$, hence also of $2 d^{\prime}$, we may write $2 d^{\prime}=m^{\prime} d^{\prime \prime}$ ( $m^{\prime}$ an integer). Since $d^{\prime \prime} \leq a_{n-2}<d^{\prime}$, we get $m^{\prime} \geq 3$. Also, $a_{n-2}<d^{\prime}=\frac{1}{2} m^{\prime} d^{\prime \prime}, a_{n+1}=\frac{15}{2} d^{\prime}=\frac{15}{4} m^{\prime} d^{\prime \prime}$. As before, we consider the cases:

$$
\begin{aligned}
& \text { if } m^{\prime} \geq 5 \quad \text { then } a_{n+1}=\frac{15}{4} m^{\prime} d^{\prime \prime} \geq \frac{75}{4} d^{\prime \prime}>\frac{75}{4} a_{n-3} \geq \frac{75}{4} \cdot 2^{n-3}>2^{n+1} ; \\
& \text { if } 3 \leq m^{\prime} \leq 4 \text { then } a_{n-2}<\frac{1}{2} \cdot 4 d^{\prime \prime}, \text { and hence } a_{n-2}=d^{\prime \prime}, \\
& \qquad a_{n+1}=\frac{15}{4} m^{\prime} a_{n-2} \geq \frac{15}{4} \cdot 3 a_{n-2} \geq \frac{45}{4} \cdot 2^{n-2}>2^{n+1} .
\end{aligned}
$$

Both of them have produced the induction claim. But now there are no cases left. Induction is complete; the inequality $a_{n} \geq 2^{n}$ holds for all $n$.
$\mathbf{N} 4$. Let $n$ be a positive integer. Show that the numbers

$$
\binom{2^{n}-1}{0}, \quad\binom{2^{n}-1}{1}, \quad\binom{2^{n}-1}{2}, \quad \ldots, \quad\binom{2^{n}-1}{2^{n-1}-1}
$$

are congruent modulo $2^{n}$ to $1,3,5, \ldots, 2^{n}-1$ in some order.
Solution 1. It is well-known that all these numbers are odd. So the assertion that their remainders $\left(\bmod 2^{n}\right)$ make up a permutation of $\left\{1,3, \ldots, 2^{n}-1\right\}$ is equivalent just to saying that these remainders are all distinct. We begin by showing that

$$
\begin{equation*}
\binom{2^{n}-1}{2 k}+\binom{2^{n}-1}{2 k+1} \equiv 0\left(\bmod 2^{n}\right) \quad \text { and } \quad\binom{2^{n}-1}{2 k} \equiv(-1)^{k}\binom{2^{n-1}-1}{k} \quad\left(\bmod 2^{n}\right) \tag{1}
\end{equation*}
$$

The first relation is immediate, as the sum on the left is equal to $\binom{2^{n}}{2 k+1}=\frac{2^{n}}{2 k+1}\binom{2^{n}-1}{2 k}$, hence is divisible by $2^{n}$. The second relation:

$$
\binom{2^{n}-1}{2 k}=\prod_{j=1}^{2 k} \frac{2^{n}-j}{j}=\prod_{i=1}^{k} \frac{2^{n}-(2 i-1)}{2 i-1} \cdot \prod_{i=1}^{k} \frac{2^{n-1}-i}{i} \equiv(-1)^{k}\binom{2^{n-1}-1}{k} \quad\left(\bmod 2^{n}\right)
$$

This prepares ground for a proof of the required result by induction on $n$. The base case $n=1$ is obvious. Assume the assertion is true for $n-1$ and pass to $n$, denoting $a_{k}=\binom{2^{n-1}-1}{k}$, $b_{m}=\binom{2^{n}-1}{m}$. The induction hypothesis is that all the numbers $a_{k}\left(0 \leq k<2^{n-2}\right)$ are distinct $\left(\bmod 2^{2^{m-1}}\right)$; the claim is that all the numbers $b_{m}\left(0 \leq m<2^{n-1}\right)$ are distinct $\left(\bmod 2^{n}\right)$.

The congruence relations (1) are restated as

$$
\begin{equation*}
b_{2 k} \equiv(-1)^{k} a_{k} \equiv-b_{2 k+1} \quad\left(\bmod 2^{n}\right) \tag{2}
\end{equation*}
$$

Shifting the exponent in the first relation of (1) from $n$ to $n-1$ we also have the congruence $a_{2 i+1} \equiv-a_{2 i}\left(\bmod 2^{n-1}\right)$. We hence conclude:

If, for some $j, k<2^{n-2}, a_{k} \equiv-a_{j}\left(\bmod 2^{n-1}\right)$, then $\{j, k\}=\{2 i, 2 i+1\}$ for some $i$.
This is so because in the sequence $\left(a_{k}: k<2^{n-2}\right)$ each term $a_{j}$ is complemented to $0\left(\bmod 2^{n-1}\right)$ by only one other term $a_{k}$, according to the induction hypothesis.

From (2) we see that $b_{4 i} \equiv a_{2 i}$ and $b_{4 i+3} \equiv a_{2 i+1}\left(\bmod 2^{n}\right)$. Let

$$
M=\left\{m: 0 \leq m<2^{n-1}, m \equiv 0 \text { or } 3(\bmod 4)\right\}, \quad L=\left\{l: 0 \leq l<2^{n-1}, l \equiv 1 \text { or } 2(\bmod 4)\right\}
$$

The last two congruences take on the unified form

$$
\begin{equation*}
b_{m} \equiv a_{\lfloor m / 2\rfloor} \quad\left(\bmod 2^{n}\right) \quad \text { for all } \quad m \in M \tag{4}
\end{equation*}
$$

Thus all the numbers $b_{m}$ for $m \in M$ are distinct $\left(\bmod 2^{n}\right)$ because so are the numbers $a_{k}$ (they are distinct $\left(\bmod 2^{n-1}\right)$, hence also $\left(\bmod 2^{n}\right)$ ).

Every $l \in L$ is paired with a unique $m \in M$ into a pair of the form $\{2 k, 2 k+1\}$. So (2) implies that also all the $b_{l}$ for $l \in L$ are distinct $\left(\bmod 2^{n}\right)$. It remains to eliminate the possibility that $b_{m} \equiv b_{l}\left(\bmod 2^{n}\right)$ for some $m \in M, l \in L$.

Suppose that such a situation occurs. Let $m^{\prime} \in M$ be such that $\left\{m^{\prime}, l\right\}$ is a pair of the form $\{2 k, 2 k+1\}$, so that $($ see $(2)) b_{m^{\prime}} \equiv-b_{l}\left(\bmod 2^{n}\right)$. Hence $b_{m^{\prime}} \equiv-b_{m}\left(\bmod 2^{n}\right)$. Since both $m^{\prime}$ and $m$ are in $M$, we have by (4) $b_{m^{\prime}} \equiv a_{j}, b_{m} \equiv a_{k}\left(\bmod 2^{n}\right)$ for $j=\left\lfloor m^{\prime} / 2\right\rfloor, k=\lfloor m / 2\rfloor$.

Then $a_{j} \equiv-a_{k}\left(\bmod 2^{n}\right)$. Thus, according to (3), $j=2 i, k=2 i+1$ for some $i$ (or vice versa). The equality $a_{2 i+1} \equiv-a_{2 i}\left(\bmod 2^{n}\right)$ now means that $\binom{2^{n-1}-1}{2 i}+\binom{2^{n-1}-1}{2 i+1} \equiv 0\left(\bmod 2^{n}\right)$. However, the sum on the left is equal to $\binom{2^{n-1}}{2 i+1}$. A number of this form cannot be divisible by $2^{n}$. This is a contradiction which concludes the induction step and proves the result.

Solution 2. We again proceed by induction, writing for brevity $N=2^{n-1}$ and keeping notation $a_{k}=\binom{N-1}{k}, b_{m}=\binom{2 N-1}{m}$. Assume that the result holds for the sequence $\left(a_{0}, a_{1}, a_{2}, \ldots, a_{N / 2-1}\right)$. In view of the symmetry $a_{N-1-k}=a_{k}$ this sequence is a permutation of ( $a_{0}, a_{2}, a_{4}, \ldots, a_{N-2}$ ). So the induction hypothesis says that this latter sequence, $\operatorname{taken}(\bmod N)$, is a permutation of $(1,3,5, \ldots, N-1)$. Similarly, the induction claim is that $\left(b_{0}, b_{2}, b_{4}, \ldots, b_{2 N-2}\right)$, taken $(\bmod 2 N)$, is a permutation of $(1,3,5, \ldots, 2 N-1)$.

In place of the congruence relations (2) we now use the following ones,

$$
\begin{equation*}
b_{4 i} \equiv a_{2 i} \quad(\bmod N) \quad \text { and } \quad b_{4 i+2} \equiv b_{4 i}+N \quad(\bmod 2 N) \tag{5}
\end{equation*}
$$

Given this, the conclusion is immediate: the first formula of (5) together with the induction hypothesis tells us that $\left(b_{0}, b_{4}, b_{8}, \ldots, b_{2 N-4}\right)(\bmod N)$ is a permutation of $(1,3,5, \ldots, N-1)$. Then the second formula of (5) shows that $\left(b_{2}, b_{6}, b_{10}, \ldots, b_{2 N-2}\right)(\bmod N)$ is exactly the same permutation; moreover, this formula distinguishes $(\bmod 2 N)$ each $b_{4 i}$ from $b_{4 i+2}$.

Consequently, these two sequences combined represent $(\bmod 2 N)$ a permutation of the sequence $(1,3,5, \ldots, N-1, N+1, N+3, N+5, \ldots, N+N-1)$, and this is precisely the induction claim.

Now we prove formulas (5); we begin with the second one. Since $b_{m+1}=b_{m} \cdot \frac{2 N-m-1}{m+1}$,

$$
b_{4 i+2}=b_{4 i} \cdot \frac{2 N-4 i-1}{4 i+1} \cdot \frac{2 N-4 i-2}{4 i+2}=b_{4 i} \cdot \frac{2 N-4 i-1}{4 i+1} \cdot \frac{N-2 i-1}{2 i+1} .
$$

The desired congruence $b_{4 i+2} \equiv b_{4 i}+N$ may be multiplied by the odd number $(4 i+1)(2 i+1)$, giving rise to a chain of successively equivalent congruences:

$$
\begin{array}{rlrl}
b_{4 i}(2 N-4 i-1)(N-2 i-1) & \equiv\left(b_{4 i}+N\right)(4 i+1)(2 i+1) & (\bmod 2 N), \\
b_{4 i}(2 i+1-N) & \equiv\left(b_{4 i}+N\right)(2 i+1) & & (\bmod 2 N), \\
\left(b_{4 i}+2 i+1\right) N & \equiv 0 & & (\bmod 2 N) ;
\end{array}
$$

and the last one is satisfied, as $b_{4 i}$ is odd. This settles the second relation in (5).
The first one is proved by induction on $i$. It holds for $i=0$. Assume $b_{4 i} \equiv a_{2 i}(\bmod 2 N)$ and consider $i+1$ :

$$
b_{4 i+4}=b_{4 i+2} \cdot \frac{2 N-4 i-3}{4 i+3} \cdot \frac{2 N-4 i-4}{4 i+4} ; \quad a_{2 i+2}=a_{2 i} \cdot \frac{N-2 i-1}{2 i+1} \cdot \frac{N-2 i-2}{2 i+2} .
$$

Both expressions have the fraction $\frac{N-2 i-2}{2 i+2}$ as the last factor. Since $2 i+2<N=2^{n-1}$, this fraction reduces to $\ell / m$ with $\ell$ and $m$ odd. In showing that $b_{4 i+4} \equiv a_{2 i+2}(\bmod 2 N)$, we may ignore this common factor $\ell / m$. Clearing other odd denominators reduces the claim to

$$
b_{4 i+2}(2 N-4 i-3)(2 i+1) \equiv a_{2 i}(N-2 i-1)(4 i+3) \quad(\bmod 2 N) .
$$

By the inductive assumption (saying that $b_{4 i} \equiv a_{2 i}(\bmod 2 N)$ ) and by the second relation of (5), this is equivalent to

$$
\left(b_{4 i}+N\right)(2 i+1) \equiv b_{4 i}(2 i+1-N) \quad(\bmod 2 N)
$$

a congruence which we have already met in the preceding proof a few lines above. This completes induction (on $i$ ) and the proof of (5), hence also the whole solution.

Comment. One can avoid the words congruent modulo in the problem statement by rephrasing the assertion into: Show that these numbers leave distinct remainders in division by $2^{n}$.

N5. For every $n \in \mathbb{N}$ let $d(n)$ denote the number of (positive) divisors of $n$. Find all functions $f: \mathbb{N} \rightarrow \mathbb{N}$ with the following properties:
(i) $d(f(x))=x$ for all $x \in \mathbb{N}$;
(ii) $f(x y)$ divides $(x-1) y^{x y-1} f(x)$ for all $x, y \in \mathbb{N}$.

Solution. There is a unique solution: the function $f: \mathbb{N} \rightarrow \mathbb{N}$ defined by $f(1)=1$ and

$$
\begin{equation*}
f(n)=p_{1}^{p_{1}^{a_{1}}-1} p_{2}^{p_{2}^{a_{2}}-1} \cdots p_{k}^{p_{k}^{a_{k}}-1} \text { where } n=p_{1}^{a_{1}} p_{2}^{a_{2}} \cdots p_{k}^{a_{k}} \text { is the prime factorization of } n>1 \tag{1}
\end{equation*}
$$

Direct verification shows that this function meets the requirements.
Conversely, let $f: \mathbb{N} \rightarrow \mathbb{N}$ satisfy (i) and (ii). Applying (i) for $x=1$ gives $d(f(1))=1$, so $f(1)=1$. In the sequel we prove that (1) holds for all $n>1$. Notice that $f(m)=f(n)$ implies $m=n$ in view of (i). The formula $d\left(p_{1}^{b_{1}} \cdots p_{k}^{b_{k}}\right)=\left(b_{1}+1\right) \cdots\left(b_{k}+1\right)$ will be used throughout.

Let $p$ be a prime. Since $d(f(p))=p$, the formula just mentioned yields $f(p)=q^{p-1}$ for some prime $q$; in particular $f(2)=q^{2-1}=q$ is a prime. We prove that $f(p)=p^{p-1}$ for all primes $p$.

Suppose that $p$ is odd and $f(p)=q^{p-1}$ for a prime $q$. Applying (ii) first with $x=2$, $y=p$ and then with $x=p, y=2$ shows that $f(2 p)$ divides both $(2-1) p^{2 p-1} f(2)=p^{2 p-1} f(2)$ and $(p-1) 2^{2 p-1} f(p)=(p-1) 2^{2 p-1} q^{p-1}$. If $q \neq p$ then the odd prime $p$ does not divide $(p-1) 2^{2 p-1} q^{p-1}$, hence the greatest common divisor of $p^{2 p-1} f(2)$ and $(p-1) 2^{2 p-1} q^{p-1}$ is a divisor of $f(2)$. Thus $f(2 p)$ divides $f(2)$ which is a prime. As $f(2 p)>1$, we obtain $f(2 p)=f(2)$ which is impossible. So $q=p$, i. e. $f(p)=p^{p-1}$.

For $p=2$ the same argument with $x=2, y=3$ and $x=3, y=2$ shows that $f(6)$ divides both $3^{5} f(2)$ and $2^{6} f(3)=2^{6} 3^{2}$. If the prime $f(2)$ is odd then $f(6)$ divides $3^{2}=9$, so $f(6) \in\{1,3,9\}$. However then $6=d(f(6)) \in\{d(1), d(3), d(9)\}=\{1,2,3\}$ which is false. In conclusion $f(2)=2$.

Next, for each $n>1$ the prime divisors of $f(n)$ are among the ones of $n$. Indeed, let $p$ be the least prime divisor of $n$. Apply (ii) with $x=p$ and $y=n / p$ to obtain that $f(n)$ divides $(p-1) y^{n-1} f(p)=(p-1) y^{n-1} p^{p-1}$. Write $f(n)=\ell P$ where $\ell$ is coprime to $n$ and $P$ is a product of primes dividing $n$. Since $\ell$ divides $(p-1) y^{n-1} p^{p-1}$ and is coprime to $y^{n-1} p^{p-1}$, it divides $p-1$; hence $d(\ell) \leq \ell<p$. But (i) gives $n=d(f(n))=d(\ell P)$, and $d(\ell P)=d(\ell) d(P)$ as $\ell$ and $P$ are coprime. Therefore $d(\ell)$ is a divisor of $n$ less than $p$, meaning that $\ell=1$ and proving the claim.

Now (1) is immediate for prime powers. If $p$ is a prime and $a \geq 1$, by the above the only prime factor of $f\left(p^{a}\right)$ is $p$ (a prime factor does exist as $f\left(p^{a}\right)>1$ ). So $f\left(p^{a}\right)=p^{b}$ for some $b \geq 1$, and (i) yields $p^{a}=d\left(f\left(p^{a}\right)\right)=d\left(p^{b}\right)=b+1$. Hence $f\left(p^{a}\right)=p^{p^{a}-1}$, as needed.

Let us finally show that ( 1 ) is true for a general $n>1$ with prime factorization $n=p_{1}^{a_{1}} \cdots p_{k}^{a_{k}}$. We saw that the prime factorization of $f(n)$ has the form $f(n)=p_{1}^{b_{1}} \cdots p_{k}^{b_{k}}$. For $i=1, \ldots, k$, set $x=p_{i}^{a_{i}}$ and $y=n / x$ in (ii) to infer that $f(n)$ divides $\left(p_{i}^{a_{i}}-1\right) y^{n-1} f\left(p_{i}^{a_{i}}\right)$. Hence $p_{i}^{b_{i}}$ divides $\left(p_{i}^{a_{i}}-1\right) y^{n-1} f\left(p_{i}^{a_{i}}\right)$, and because $p_{i}^{b_{i}}$ is coprime to $\left(p_{i}^{a_{i}}-1\right) y^{n-1}$, it follows that $p_{i}^{b_{i}}$ divides $f\left(p_{i}^{a_{i}}\right)=p_{i}^{p_{i}^{a_{i}}-1}$. So $b_{i} \leq p_{i}^{a_{i}}-1$ for all $i=1, \ldots, k$. Combined with (i), these conclusions imply

$$
p_{1}^{a_{1}} \cdots p_{k}^{a_{k}}=n=d(f(n))=d\left(p_{1}^{b_{1}} \cdots p_{k}^{b_{k}}\right)=\left(b_{1}+1\right) \cdots\left(b_{k}+1\right) \leq p_{1}^{a_{1}} \cdots p_{k}^{a_{k}}
$$

Hence all inequalities $b_{i} \leq p_{i}^{a_{i}}-1$ must be equalities, $i=1, \ldots, k$, implying that (1) holds true. The proof is complete.

N6. Prove that there exist infinitely many positive integers $n$ such that $n^{2}+1$ has a prime divisor greater than $2 n+\sqrt{2 n}$.

Solution. Let $p \equiv 1(\bmod 8)$ be a prime. The congruence $x^{2} \equiv-1(\bmod p)$ has two solutions in $[1, p-1]$ whose sum is $p$. If $n$ is the smaller one of them then $p$ divides $n^{2}+1$ and $n \leq(p-1) / 2$. We show that $p>2 n+\sqrt{10 n}$.

Let $n=(p-1) / 2-\ell$ where $\ell \geq 0$. Then $n^{2} \equiv-1(\bmod p)$ gives

$$
\left(\frac{p-1}{2}-\ell\right)^{2} \equiv-1 \quad(\bmod p) \quad \text { or } \quad(2 \ell+1)^{2}+4 \equiv 0 \quad(\bmod p)
$$

Thus $(2 \ell+1)^{2}+4=r p$ for some $r \geq 0$. As $(2 \ell+1)^{2} \equiv 1 \equiv p(\bmod 8)$, we have $r \equiv 5(\bmod 8)$, so that $r \geq 5$. Hence $(2 \ell+1)^{2}+4 \geq 5 p$, implying $\ell \geq(\sqrt{5 p-4}-1) / 2$. Set $\sqrt{5 p-4}=u$ for clarity; then $\ell \geq(u-1) / 2$. Therefore

$$
n=\frac{p-1}{2}-\ell \leq \frac{1}{2}(p-u) .
$$

Combined with $p=\left(u^{2}+4\right) / 5$, this leads to $u^{2}-5 u-10 n+4 \geq 0$. Solving this quadratic inequality with respect to $u \geq 0$ gives $u \geq(5+\sqrt{40 n+9}) / 2$. So the estimate $n \leq(p-u) / 2$ leads to

$$
p \geq 2 n+u \geq 2 n+\frac{1}{2}(5+\sqrt{40 n+9})>2 n+\sqrt{10 n}
$$

Since there are infinitely many primes of the form $8 k+1$, it follows easily that there are also infinitely many $n$ with the stated property.

Comment. By considering the prime factorization of the product $\prod_{n=1}^{N}\left(n^{2}+1\right)$, it can be obtained that its greatest prime divisor is at least $c N \log N$. This could improve the statement as $p>n \log n$.

However, the proof applies some advanced information about the distribution of the primes of the form $4 k+1$, which is inappropriate for high schools contests.


## International Mathematical Olympiad

 Bremen Germany 2009
## 10 to 22 July 2009

## Problem Shortist with solutions



## Problem Shortlist with Solutions

The Problem Selection Committee

We insistently ask everybody to consider the following IMO Regulations rule:

## These Shortlist Problems have to be kept strictly confidential until IMO 2010.

## The Problem Selection Committee

Konrad Engel, Karl Fegert, Andreas Felgenhauer, Hans-Dietrich Gronau, Roger Labahn, Bernd Mulansky, Jürgen Prestin, Christian Reiher, Peter Scholze, Eckard Specht, Robert Strich, Martin Welk
gratefully received
132 problem proposals submitted by 39 countries:
Algeria, Australia, Austria, Belarus, Belgium, Bulgaria, Colombia, Croatia, Czech Republic, El Salvador, Estonia, Finland, France, Greece, Hong Kong, Hungary, India, Ireland, Islamic Republic of Iran, Japan, Democratic People's Republic of Korea, Lithuania, Luxembourg, The former Yugoslav Republic of Macedonia, Mongolia, Netherlands, New Zealand, Pakistan, Peru, Poland, Romania, Russian Federation, Slovenia, South Africa, Taiwan, Turkey, Ukraine, United Kingdom, United States of America.

Layout: Roger Labahn with $\mathrm{ET}_{\mathrm{E}} \mathrm{X} \& \mathrm{~T}_{\mathrm{E}} \mathrm{X}$
Drawings: Eckard Specht with nicefig 2.0


The Problem Selection Committee

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## Algebra

## A1 CZE (Czech Republic)

Find the largest possible integer $k$, such that the following statement is true:
Let 2009 arbitrary non-degenerated triangles be given. In every triangle the three sides are colored, such that one is blue, one is red and one is white. Now, for every color separately, let us sort the lengths of the sides. We obtain

$$
\begin{array}{rlrl}
b_{1} & \leq b_{2} & \leq \ldots \leq b_{2009} & \\
& & \text { the lengths of the blue sides, } \\
r_{1} & \leq r_{2} & \leq \ldots \leq r_{2009} & \text { the lengths of the red sides, } \\
\text { and } \quad w_{1} & \leq w_{2} \leq \ldots \leq w_{2009} \quad \text { the lengths of the white sides. }
\end{array}
$$

Then there exist $k$ indices $j$ such that we can form a non-degenerated triangle with side lengths $b_{j}, r_{j}, w_{j}$.

## A2 EST (Estonia)

Let $a, b, c$ be positive real numbers such that $\frac{1}{a}+\frac{1}{b}+\frac{1}{c}=a+b+c$. Prove that

$$
\frac{1}{(2 a+b+c)^{2}}+\frac{1}{(2 b+c+a)^{2}}+\frac{1}{(2 c+a+b)^{2}} \leq \frac{3}{16} .
$$

## A3 FRA (France)

Determine all functions $f$ from the set of positive integers into the set of positive integers such that for all $x$ and $y$ there exists a non degenerated triangle with sides of lengths

$$
x, \quad f(y) \quad \text { and } \quad f(y+f(x)-1) .
$$

## A4 BLR (Belarus)

Let $a, b, c$ be positive real numbers such that $a b+b c+c a \leq 3 a b c$. Prove that

$$
\sqrt{\frac{a^{2}+b^{2}}{a+b}}+\sqrt{\frac{b^{2}+c^{2}}{b+c}}+\sqrt{\frac{c^{2}+a^{2}}{c+a}}+3 \leq \sqrt{2}(\sqrt{a+b}+\sqrt{b+c}+\sqrt{c+a}) .
$$

## A5 BLR (Belarus)

Let $f$ be any function that maps the set of real numbers into the set of real numbers. Prove that there exist real numbers $x$ and $y$ such that

$$
f(x-f(y))>y f(x)+x .
$$

## A6 USA (United States of America)

Suppose that $s_{1}, s_{2}, s_{3}, \ldots$ is a strictly increasing sequence of positive integers such that the subsequences

$$
s_{s_{1}}, s_{s_{2}}, s_{s_{3}}, \ldots \quad \text { and } \quad s_{s_{1}+1}, s_{s_{2}+1}, s_{s_{3}+1}, \ldots
$$

are both arithmetic progressions. Prove that $s_{1}, s_{2}, s_{3}, \ldots$ is itself an arithmetic progression.

## A7 JPN (Japan)

Find all functions $f$ from the set of real numbers into the set of real numbers which satisfy for all real $x, y$ the identity

$$
f(x f(x+y))=f(y f(x))+x^{2}
$$

## Combinatorics

## C1 NZL (New Zealand)

Consider 2009 cards, each having one gold side and one black side, lying in parallel on a long table. Initially all cards show their gold sides. Two players, standing by the same long side of the table, play a game with alternating moves. Each move consists of choosing a block of 50 consecutive cards, the leftmost of which is showing gold, and turning them all over, so those which showed gold now show black and vice versa. The last player who can make a legal move wins.
(a) Does the game necessarily end?
(b) Does there exist a winning strategy for the starting player?

## C2 ROU (Romania)

For any integer $n \geq 2$, let $N(n)$ be the maximal number of triples $\left(a_{i}, b_{i}, c_{i}\right), i=1, \ldots, N(n)$, consisting of nonnegative integers $a_{i}, b_{i}$ and $c_{i}$ such that the following two conditions are satisfied:
(1) $a_{i}+b_{i}+c_{i}=n$ for all $i=1, \ldots, N(n)$,
(2) If $i \neq j$, then $a_{i} \neq a_{j}, b_{i} \neq b_{j}$ and $c_{i} \neq c_{j}$.

Determine $N(n)$ for all $n \geq 2$.

Comment. The original problem was formulated for $m$-tuples instead for triples. The numbers $N(m, n)$ are then defined similarly to $N(n)$ in the case $m=3$. The numbers $N(3, n)$ and $N(n, n)$ should be determined. The case $m=3$ is the same as in the present problem. The upper bound for $N(n, n)$ can be proved by a simple generalization. The construction of a set of triples attaining the bound can be easily done by induction from $n$ to $n+2$.

## C3 RUS (Russian Federation)

Let $n$ be a positive integer. Given a sequence $\varepsilon_{1}, \ldots, \varepsilon_{n-1}$ with $\varepsilon_{i}=0$ or $\varepsilon_{i}=1$ for each $i=1, \ldots, n-1$, the sequences $a_{0}, \ldots, a_{n}$ and $b_{0}, \ldots, b_{n}$ are constructed by the following rules:

$$
\begin{gathered}
a_{0}=b_{0}=1, \quad a_{1}=b_{1}=7, \\
a_{i+1}=\left\{\begin{array}{ll}
2 a_{i-1}+3 a_{i}, & \text { if } \varepsilon_{i}=0, \\
3 a_{i-1}+a_{i}, & \text { if } \varepsilon_{i}=1,
\end{array} \text { for each } i=1, \ldots, n-1,\right. \\
b_{i+1}=\left\{\begin{array}{ll}
2 b_{i-1}+3 b_{i}, & \text { if } \varepsilon_{n-i}=0, \\
3 b_{i-1}+b_{i}, & \text { if } \varepsilon_{n-i}=1,
\end{array} \text { for each } i=1, \ldots, n-1 .\right.
\end{gathered}
$$

Prove that $a_{n}=b_{n}$.

## C4 NLD (Netherlands)

For an integer $m \geq 1$, we consider partitions of a $2^{m} \times 2^{m}$ chessboard into rectangles consisting of cells of the chessboard, in which each of the $2^{m}$ cells along one diagonal forms a separate rectangle of side length 1 . Determine the smallest possible sum of rectangle perimeters in such a partition.

## C5 NLD (Netherlands)

Five identical empty buckets of 2-liter capacity stand at the vertices of a regular pentagon. Cinderella and her wicked Stepmother go through a sequence of rounds: At the beginning of every round, the Stepmother takes one liter of water from the nearby river and distributes it arbitrarily over the five buckets. Then Cinderella chooses a pair of neighboring buckets, empties them into the river, and puts them back. Then the next round begins. The Stepmother's goal is to make one of these buckets overflow. Cinderella's goal is to prevent this. Can the wicked Stepmother enforce a bucket overflow?

## C6 BGR (Bulgaria)

On a $999 \times 999$ board a limp rook can move in the following way: From any square it can move to any of its adjacent squares, i.e. a square having a common side with it, and every move must be a turn, i.e. the directions of any two consecutive moves must be perpendicular. A nonintersecting route of the limp rook consists of a sequence of pairwise different squares that the limp rook can visit in that order by an admissible sequence of moves. Such a non-intersecting route is called cyclic, if the limp rook can, after reaching the last square of the route, move directly to the first square of the route and start over.
How many squares does the longest possible cyclic, non-intersecting route of a limp rook visit?

## C7 RUS (Russian Federation)

Variant 1. A grasshopper jumps along the real axis. He starts at point 0 and makes 2009 jumps to the right with lengths $1,2, \ldots, 2009$ in an arbitrary order. Let $M$ be a set of 2008 positive integers less than $1005 \cdot 2009$. Prove that the grasshopper can arrange his jumps in such a way that he never lands on a point from $M$.

Variant 2. Let $n$ be a nonnegative integer. A grasshopper jumps along the real axis. He starts at point 0 and makes $n+1$ jumps to the right with pairwise different positive integral lengths $a_{1}, a_{2}, \ldots, a_{n+1}$ in an arbitrary order. Let $M$ be a set of $n$ positive integers in the interval $(0, s)$, where $s=a_{1}+a_{2}+\cdots+a_{n+1}$. Prove that the grasshopper can arrange his jumps in such a way that he never lands on a point from $M$.

## C8 AUT (Austria)

For any integer $n \geq 2$, we compute the integer $h(n)$ by applying the following procedure to its decimal representation. Let $r$ be the rightmost digit of $n$.
(1) If $r=0$, then the decimal representation of $h(n)$ results from the decimal representation of $n$ by removing this rightmost digit 0 .
(2) If $1 \leq r \leq 9$ we split the decimal representation of $n$ into a maximal right part $R$ that solely consists of digits not less than $r$ and into a left part $L$ that either is empty or ends with a digit strictly smaller than $r$. Then the decimal representation of $h(n)$ consists of the decimal representation of $L$, followed by two copies of the decimal representation of $R-1$. For instance, for the number $n=17,151,345,543$, we will have $L=17,151, R=345,543$ and $h(n)=17,151,345,542,345,542$.
Prove that, starting with an arbitrary integer $n \geq 2$, iterated application of $h$ produces the integer 1 after finitely many steps.

## Geometry

## G1 BEL (Belgium)

Let $A B C$ be a triangle with $A B=A C$. The angle bisectors of $A$ and $B$ meet the sides $B C$ and $A C$ in $D$ and $E$, respectively. Let $K$ be the incenter of triangle $A D C$. Suppose that $\angle B E K=45^{\circ}$. Find all possible values of $\angle B A C$.

## G2 RUS (Russian Federation)

Let $A B C$ be a triangle with circumcenter $O$. The points $P$ and $Q$ are interior points of the sides $C A$ and $A B$, respectively. The circle $k$ passes through the midpoints of the segments $B P$, $C Q$, and $P Q$. Prove that if the line $P Q$ is tangent to circle $k$ then $O P=O Q$.

## G3 IRN (Islamic Republic of Iran)

Let $A B C$ be a triangle. The incircle of $A B C$ touches the sides $A B$ and $A C$ at the points $Z$ and $Y$, respectively. Let $G$ be the point where the lines $B Y$ and $C Z$ meet, and let $R$ and $S$ be points such that the two quadrilaterals $B C Y R$ and $B C S Z$ are parallelograms.
Prove that $G R=G S$.

## G4 UNK (United Kingdom)

Given a cyclic quadrilateral $A B C D$, let the diagonals $A C$ and $B D$ meet at $E$ and the lines $A D$ and $B C$ meet at $F$. The midpoints of $A B$ and $C D$ are $G$ and $H$, respectively. Show that $E F$ is tangent at $E$ to the circle through the points $E, G$, and $H$.

## G5 POL (Poland)

Let $P$ be a polygon that is convex and symmetric to some point $O$. Prove that for some parallelogram $R$ satisfying $P \subset R$ we have

$$
\frac{|R|}{|P|} \leq \sqrt{2}
$$

where $|R|$ and $|P|$ denote the area of the sets $R$ and $P$, respectively.

## G6 UKR (Ukraine)

Let the sides $A D$ and $B C$ of the quadrilateral $A B C D$ (such that $A B$ is not parallel to $C D$ ) intersect at point $P$. Points $O_{1}$ and $O_{2}$ are the circumcenters and points $H_{1}$ and $H_{2}$ are the orthocenters of triangles $A B P$ and $D C P$, respectively. Denote the midpoints of segments $O_{1} H_{1}$ and $O_{2} H_{2}$ by $E_{1}$ and $E_{2}$, respectively. Prove that the perpendicular from $E_{1}$ on $C D$, the perpendicular from $E_{2}$ on $A B$ and the line $H_{1} H_{2}$ are concurrent.

## G7 IRN (Islamic Republic of Iran)

Let $A B C$ be a triangle with incenter $I$ and let $X, Y$ and $Z$ be the incenters of the triangles $B I C, C I A$ and $A I B$, respectively. Let the triangle $X Y Z$ be equilateral. Prove that $A B C$ is equilateral too.

## G8 BGR (Bulgaria)

Let $A B C D$ be a circumscribed quadrilateral. Let $g$ be a line through $A$ which meets the segment $B C$ in $M$ and the line $C D$ in $N$. Denote by $I_{1}, I_{2}$, and $I_{3}$ the incenters of $\triangle A B M$, $\triangle M N C$, and $\triangle N D A$, respectively. Show that the orthocenter of $\triangle I_{1} I_{2} I_{3}$ lies on $g$.

## Number Theory

## N1 AUS (Australia)

A social club has $n$ members. They have the membership numbers $1,2, \ldots, n$, respectively. From time to time members send presents to other members, including items they have already received as presents from other members. In order to avoid the embarrassing situation that a member might receive a present that he or she has sent to other members, the club adds the following rule to its statutes at one of its annual general meetings:
"A member with membership number $a$ is permitted to send a present to a member with membership number $b$ if and only if $a(b-1)$ is a multiple of $n$."
Prove that, if each member follows this rule, none will receive a present from another member that he or she has already sent to other members.

Alternative formulation: Let $G$ be a directed graph with $n$ vertices $v_{1}, v_{2}, \ldots, v_{n}$, such that there is an edge going from $v_{a}$ to $v_{b}$ if and only if $a$ and $b$ are distinct and $a(b-1)$ is a multiple of $n$. Prove that this graph does not contain a directed cycle.

## N2 PER (Peru)

A positive integer $N$ is called balanced, if $N=1$ or if $N$ can be written as a product of an even number of not necessarily distinct primes. Given positive integers $a$ and $b$, consider the polynomial $P$ defined by $P(x)=(x+a)(x+b)$.
(a) Prove that there exist distinct positive integers $a$ and $b$ such that all the numbers $P(1), P(2)$, $\ldots, P(50)$ are balanced.
(b) Prove that if $P(n)$ is balanced for all positive integers $n$, then $a=b$.

## N3 EST (Estonia)

Let $f$ be a non-constant function from the set of positive integers into the set of positive integers, such that $a-b$ divides $f(a)-f(b)$ for all distinct positive integers $a, b$. Prove that there exist infinitely many primes $p$ such that $p$ divides $f(c)$ for some positive integer $c$.

## N4 PRK (Democratic People's Republic of Korea)

Find all positive integers $n$ such that there exists a sequence of positive integers $a_{1}, a_{2}, \ldots, a_{n}$ satisfying

$$
a_{k+1}=\frac{a_{k}^{2}+1}{a_{k-1}+1}-1
$$

for every $k$ with $2 \leq k \leq n-1$.

## N5 HUN (Hungary)

Let $P(x)$ be a non-constant polynomial with integer coefficients. Prove that there is no function $T$ from the set of integers into the set of integers such that the number of integers $x$ with $T^{n}(x)=x$ is equal to $P(n)$ for every $n \geq 1$, where $T^{n}$ denotes the $n$-fold application of $T$.

## N6 TUR (Turkey)

Let $k$ be a positive integer. Show that if there exists a sequence $a_{0}, a_{1}, \ldots$ of integers satisfying the condition

$$
a_{n}=\frac{a_{n-1}+n^{k}}{n} \quad \text { for all } n \geq 1,
$$

then $k-2$ is divisible by 3 .

## N7 MNG (Mongolia)

Let $a$ and $b$ be distinct integers greater than 1 . Prove that there exists a positive integer $n$ such that $\left(a^{n}-1\right)\left(b^{n}-1\right)$ is not a perfect square.

## Algebra

## A1 CZE (Czech Republic)

Find the largest possible integer $k$, such that the following statement is true:
Let 2009 arbitrary non-degenerated triangles be given. In every triangle the three sides are colored, such that one is blue, one is red and one is white. Now, for every color separately, let us sort the lengths of the sides. We obtain

$$
\begin{array}{rlrl} 
& b_{1} & \leq b_{2} & \leq \ldots \leq b_{2009} \\
& & \text { the lengths of the blue sides, } \\
r_{1} & \leq r_{2} \leq \ldots \leq r_{2009} & \text { the lengths of the red sides, } \\
\text { and } \quad w_{1} & \leq w_{2} \leq \ldots \leq w_{2009} & & \text { the lengths of the white sides. }
\end{array}
$$

Then there exist $k$ indices $j$ such that we can form a non-degenerated triangle with side lengths $b_{j}, r_{j}, w_{j}$.

Solution. We will prove that the largest possible number $k$ of indices satisfying the given condition is one.

Firstly we prove that $b_{2009}, r_{2009}, w_{2009}$ are always lengths of the sides of a triangle. Without loss of generality we may assume that $w_{2009} \geq r_{2009} \geq b_{2009}$. We show that the inequality $b_{2009}+r_{2009}>w_{2009}$ holds. Evidently, there exists a triangle with side lengths $w, b, r$ for the white, blue and red side, respectively, such that $w_{2009}=w$. By the conditions of the problem we have $b+r>w, b_{2009} \geq b$ and $r_{2009} \geq r$. From these inequalities it follows

$$
b_{2009}+r_{2009} \geq b+r>w=w_{2009} .
$$

Secondly we will describe a sequence of triangles for which $w_{j}, b_{j}, r_{j}$ with $j<2009$ are not the lengths of the sides of a triangle. Let us define the sequence $\Delta_{j}, j=1,2, \ldots, 2009$, of triangles, where $\Delta_{j}$ has
a blue side of length $2 j$,
a red side of length $j$ for all $j \leq 2008$ and 4018 for $j=2009$,
and a white side of length $j+1$ for all $j \leq 2007,4018$ for $j=2008$ and 1 for $j=2009$.
Since

$$
\begin{array}{rlll}
(j+1)+j>2 j & \geq j+1>j, & \text { if } \quad j \leq 2007, \\
2 j+j>4018>2 j \quad>j, & \text { if } \quad j=2008 \\
4018+1>2 j & =4018>1, & \text { if } & j=2009
\end{array}
$$

such a sequence of triangles exists. Moreover, $w_{j}=j, r_{j}=j$ and $b_{j}=2 j$ for $1 \leq j \leq 2008$. Then

$$
w_{j}+r_{j}=j+j=2 j=b_{j},
$$

i.e., $b_{j}, r_{j}$ and $w_{j}$ are not the lengths of the sides of a triangle for $1 \leq j \leq 2008$.

## A2 EST (Estonia)

Let $a, b, c$ be positive real numbers such that $\frac{1}{a}+\frac{1}{b}+\frac{1}{c}=a+b+c$. Prove that

$$
\frac{1}{(2 a+b+c)^{2}}+\frac{1}{(2 b+c+a)^{2}}+\frac{1}{(2 c+a+b)^{2}} \leq \frac{3}{16}
$$

Solution 1. For positive real numbers $x, y, z$, from the arithmetic-geometric-mean inequality,

$$
2 x+y+z=(x+y)+(x+z) \geq 2 \sqrt{(x+y)(x+z)}
$$

we obtain

$$
\frac{1}{(2 x+y+z)^{2}} \leq \frac{1}{4(x+y)(x+z)}
$$

Applying this to the left-hand side terms of the inequality to prove, we get

$$
\begin{align*}
\frac{1}{(2 a+b+c)^{2}} & +\frac{1}{(2 b+c+a)^{2}}+\frac{1}{(2 c+a+b)^{2}} \\
& \leq \frac{1}{4(a+b)(a+c)}+\frac{1}{4(b+c)(b+a)}+\frac{1}{4(c+a)(c+b)} \\
& =\frac{(b+c)+(c+a)+(a+b)}{4(a+b)(b+c)(c+a)}=\frac{a+b+c}{2(a+b)(b+c)(c+a)} . \tag{1}
\end{align*}
$$

A second application of the inequality of the arithmetic-geometric mean yields

$$
a^{2} b+a^{2} c+b^{2} a+b^{2} c+c^{2} a+c^{2} b \geq 6 a b c
$$

or, equivalently,

$$
\begin{equation*}
9(a+b)(b+c)(c+a) \geq 8(a+b+c)(a b+b c+c a) \tag{2}
\end{equation*}
$$

The supposition $\frac{1}{a}+\frac{1}{b}+\frac{1}{c}=a+b+c$ can be written as

$$
\begin{equation*}
a b+b c+c a=a b c(a+b+c) \tag{3}
\end{equation*}
$$

Applying the arithmetic-geometric-mean inequality $x^{2} y^{2}+x^{2} z^{2} \geq 2 x^{2} y z$ thrice, we get

$$
a^{2} b^{2}+b^{2} c^{2}+c^{2} a^{2} \geq a^{2} b c+a b^{2} c+a b c^{2}
$$

which is equivalent to

$$
\begin{equation*}
(a b+b c+c a)^{2} \geq 3 a b c(a+b+c) \tag{4}
\end{equation*}
$$

Combining (1), (2), (3), and (4), we will finish the proof:

$$
\begin{aligned}
\frac{a+b+c}{2(a+b)(b+c)(c+a)} & =\frac{(a+b+c)(a b+b c+c a)}{2(a+b)(b+c)(c+a)} \cdot \frac{a b+b c+c a}{a b c(a+b+c)} \cdot \frac{a b c(a+b+c)}{(a b+b c+c a)^{2}} \\
& \leq \frac{9}{2 \cdot 8} \cdot 1 \cdot \frac{1}{3}=\frac{3}{16}
\end{aligned}
$$

Solution 2. Equivalently, we prove the homogenized inequality

$$
\frac{(a+b+c)^{2}}{(2 a+b+c)^{2}}+\frac{(a+b+c)^{2}}{(a+2 b+c)^{2}}+\frac{(a+b+c)^{2}}{(a+b+2 c)^{2}} \leq \frac{3}{16}(a+b+c)\left(\frac{1}{a}+\frac{1}{b}+\frac{1}{c}\right)
$$

for all positive real numbers $a, b, c$. Without loss of generality we choose $a+b+c=1$. Thus, the problem is equivalent to prove for all $a, b, c>0$, fulfilling this condition, the inequality

$$
\begin{equation*}
\frac{1}{(1+a)^{2}}+\frac{1}{(1+b)^{2}}+\frac{1}{(1+c)^{2}} \leq \frac{3}{16}\left(\frac{1}{a}+\frac{1}{b}+\frac{1}{c}\right) \tag{5}
\end{equation*}
$$

Applying Jensen's inequality to the function $f(x)=\frac{x}{(1+x)^{2}}$, which is concave for $0 \leq x \leq 2$ and increasing for $0 \leq x \leq 1$, we obtain

$$
\alpha \frac{a}{(1+a)^{2}}+\beta \frac{b}{(1+b)^{2}}+\gamma \frac{c}{(1+c)^{2}} \leq(\alpha+\beta+\gamma) \frac{A}{(1+A)^{2}}, \quad \text { where } \quad A=\frac{\alpha a+\beta b+\gamma c}{\alpha+\beta+\gamma} .
$$

Choosing $\alpha=\frac{1}{a}, \beta=\frac{1}{b}$, and $\gamma=\frac{1}{c}$, we can apply the harmonic-arithmetic-mean inequality

$$
A=\frac{3}{\frac{1}{a}+\frac{1}{b}+\frac{1}{c}} \leq \frac{a+b+c}{3}=\frac{1}{3}<1 .
$$

Finally we prove (5):

$$
\begin{aligned}
\frac{1}{(1+a)^{2}}+\frac{1}{(1+b)^{2}}+\frac{1}{(1+c)^{2}} & \leq\left(\frac{1}{a}+\frac{1}{b}+\frac{1}{c}\right) \frac{A}{(1+A)^{2}} \\
& \leq\left(\frac{1}{a}+\frac{1}{b}+\frac{1}{c}\right) \frac{\frac{1}{3}}{\left(1+\frac{1}{3}\right)^{2}}=\frac{3}{16}\left(\frac{1}{a}+\frac{1}{b}+\frac{1}{c}\right)
\end{aligned}
$$

## A3 FRA (France)

Determine all functions $f$ from the set of positive integers into the set of positive integers such that for all $x$ and $y$ there exists a non degenerated triangle with sides of lengths

$$
x, \quad f(y) \quad \text { and } \quad f(y+f(x)-1) .
$$

Solution. The identity function $f(x)=x$ is the only solution of the problem.
If $f(x)=x$ for all positive integers $x$, the given three lengths are $x, y=f(y)$ and $z=$ $f(y+f(x)-1)=x+y-1$. Because of $x \geq 1, y \geq 1$ we have $z \geq \max \{x, y\}>|x-y|$ and $z<x+y$. From this it follows that a triangle with these side lengths exists and does not degenerate. We prove in several steps that there is no other solution.

Step 1. We show $f(1)=1$.
If we had $f(1)=1+m>1$ we would conclude $f(y)=f(y+m)$ for all $y$ considering the triangle with the side lengths $1, f(y)$ and $f(y+m)$. Thus, $f$ would be $m$-periodic and, consequently, bounded. Let $B$ be a bound, $f(x) \leq B$. If we choose $x>2 B$ we obtain the contradiction $x>2 B \geq f(y)+f(y+f(x)-1)$.

Step 2. For all positive integers $z$, we have $f(f(z))=z$.
Setting $x=z$ and $y=1$ this follows immediately from Step 1 .

Step 3. For all integers $z \geq 1$, we have $f(z) \leq z$.
Let us show, that the contrary leads to a contradiction. Assume $w+1=f(z)>z$ for some $z$. From Step 1 we know that $w \geq z \geq 2$. Let $M=\max \{f(1), f(2), \ldots, f(w)\}$ be the largest value of $f$ for the first $w$ integers. First we show, that no positive integer $t$ exists with

$$
\begin{equation*}
f(t)>\frac{z-1}{w} \cdot t+M \tag{1}
\end{equation*}
$$

otherwise we decompose the smallest value $t$ as $t=w r+s$ where $r$ is an integer and $1 \leq s \leq w$. Because of the definition of $M$, we have $t>w$. Setting $x=z$ and $y=t-w$ we get from the triangle inequality

$$
z+f(t-w)>f((t-w)+f(z)-1)=f(t-w+w)=f(t)
$$

Hence,

$$
f(t-w) \geq f(t)-(z-1)>\frac{z-1}{w}(t-w)+M
$$

a contradiction to the minimality of $t$.
Therefore the inequality (1) fails for all $t \geq 1$, we have proven

$$
\begin{equation*}
f(t) \leq \frac{z-1}{w} \cdot t+M \tag{2}
\end{equation*}
$$

instead.

Now, using (2), we finish the proof of Step 3. Because of $z \leq w$ we have $\frac{z-1}{w}<1$ and we can choose an integer $t$ sufficiently large to fulfill the condition

$$
\left(\frac{z-1}{w}\right)^{2} t+\left(\frac{z-1}{w}+1\right) M<t .
$$

Applying (2) twice we get

$$
f(f(t)) \leq \frac{z-1}{w} f(t)+M \leq \frac{z-1}{w}\left(\frac{z-1}{w} t+M\right)+M<t
$$

in contradiction to Step 2, which proves Step 3.

Final step. Thus, following Step 2 and Step 3, we obtain

$$
z=f(f(z)) \leq f(z) \leq z
$$

and $f(z)=z$ for all positive integers $z$ is proven.

## A4 BLR (Belarus)

Let $a, b, c$ be positive real numbers such that $a b+b c+c a \leq 3 a b c$. Prove that

$$
\sqrt{\frac{a^{2}+b^{2}}{a+b}}+\sqrt{\frac{b^{2}+c^{2}}{b+c}}+\sqrt{\frac{c^{2}+a^{2}}{c+a}}+3 \leq \sqrt{2}(\sqrt{a+b}+\sqrt{b+c}+\sqrt{c+a})
$$

Solution. Starting with the terms of the right-hand side, the quadratic-arithmetic-mean inequality yields

$$
\begin{aligned}
\sqrt{2} \sqrt{a+b} & =2 \sqrt{\frac{a b}{a+b}} \sqrt{\frac{1}{2}\left(2+\frac{a^{2}+b^{2}}{a b}\right)} \\
& \geq 2 \sqrt{\frac{a b}{a+b}} \cdot \frac{1}{2}\left(\sqrt{2}+\sqrt{\frac{a^{2}+b^{2}}{a b}}\right)=\sqrt{\frac{2 a b}{a+b}}+\sqrt{\frac{a^{2}+b^{2}}{a+b}}
\end{aligned}
$$

and, analogously,

$$
\sqrt{2} \sqrt{b+c} \geq \sqrt{\frac{2 b c}{b+c}}+\sqrt{\frac{b^{2}+c^{2}}{b+c}}, \quad \sqrt{2} \sqrt{c+a} \geq \sqrt{\frac{2 c a}{c+a}}+\sqrt{\frac{c^{2}+a^{2}}{c+a}}
$$

Applying the inequality between the arithmetic mean and the squared harmonic mean will finish the proof:

$$
\sqrt{\frac{2 a b}{a+b}}+\sqrt{\frac{2 b c}{b+c}}+\sqrt{\frac{2 c a}{c+a}} \geq 3 \cdot \sqrt{\frac{3}{\sqrt{\frac{a+b}{2 a b}}^{2}+\sqrt{\frac{b+c}{2 b c}}+\sqrt{\frac{c+a}{2 c a}}}}{ }^{2}-3 \cdot \sqrt{\frac{3 a b c}{a b+b c+c a}} \geq 3
$$

## A5 BLR (Belarus)

Let $f$ be any function that maps the set of real numbers into the set of real numbers. Prove that there exist real numbers $x$ and $y$ such that

$$
f(x-f(y))>y f(x)+x .
$$

Solution 1. Assume that

$$
\begin{equation*}
f(x-f(y)) \leq y f(x)+x \quad \text { for all real } x, y \tag{1}
\end{equation*}
$$

Let $a=f(0)$. Setting $y=0$ in (11) gives $f(x-a) \leq x$ for all real $x$ and, equivalently,

$$
\begin{equation*}
f(y) \leq y+a \quad \text { for all real } y \tag{2}
\end{equation*}
$$

Setting $x=f(y)$ in (1) yields in view of (2)

$$
a=f(0) \leq y f(f(y))+f(y) \leq y f(f(y))+y+a .
$$

This implies $0 \leq y(f(f(y))+1)$ and thus

$$
\begin{equation*}
f(f(y)) \geq-1 \quad \text { for all } y>0 . \tag{3}
\end{equation*}
$$

From (2) and (3) we obtain $-1 \leq f(f(y)) \leq f(y)+a$ for all $y>0$, so

$$
\begin{equation*}
f(y) \geq-a-1 \quad \text { for all } y>0 . \tag{4}
\end{equation*}
$$

Now we show that

$$
\begin{equation*}
f(x) \leq 0 \quad \text { for all real } x . \tag{5}
\end{equation*}
$$

Assume the contrary, i.e. there is some $x$ such that $f(x)>0$. Take any $y$ such that

$$
y<x-a \quad \text { and } \quad y<\frac{-a-x-1}{f(x)} .
$$

Then in view of (2)

$$
x-f(y) \geq x-(y+a)>0
$$

and with (1) and (4) we obtain

$$
y f(x)+x \geq f(x-f(y)) \geq-a-1,
$$

whence

$$
y \geq \frac{-a-x-1}{f(x)}
$$

contrary to our choice of $y$. Thereby, we have established (5).
Setting $x=0$ in (5) leads to $a=f(0) \leq 0$ and (2) then yields

$$
\begin{equation*}
f(x) \leq x \quad \text { for all real } x \tag{6}
\end{equation*}
$$

Now choose $y$ such that $y>0$ and $y>-f(-1)-1$ and set $x=f(y)-1$. From (11), (5) and
(6) we obtain

$$
f(-1)=f(x-f(y)) \leq y f(x)+x=y f(f(y)-1)+f(y)-1 \leq y(f(y)-1)-1 \leq-y-1,
$$

i.e. $y \leq-f(-1)-1$, a contradiction to the choice of $y$.

Solution 2. Assume that

$$
\begin{equation*}
f(x-f(y)) \leq y f(x)+x \quad \text { for all real } x, y \tag{7}
\end{equation*}
$$

Let $a=f(0)$. Setting $y=0$ in (7) gives $f(x-a) \leq x$ for all real $x$ and, equivalently,

$$
\begin{equation*}
f(y) \leq y+a \quad \text { for all real } y \tag{8}
\end{equation*}
$$

Now we show that

$$
\begin{equation*}
f(z) \geq 0 \quad \text { for all } z \geq 1 \tag{9}
\end{equation*}
$$

Let $z \geq 1$ be fixed, set $b=f(z)$ and assume that $b<0$. Setting $x=w+b$ and $y=z$ in (7) gives

$$
\begin{equation*}
f(w)-z f(w+b) \leq w+b \quad \text { for all real } w \tag{10}
\end{equation*}
$$

Applying (10) to $w, w+b, \ldots, w+(n-1) b$, where $n=1,2, \ldots$, leads to

$$
\begin{aligned}
& f(w)-z^{n} f(w+n b)=(f(w)-z f(w+b))+z(f(w+b)-z f(w+2 b)) \\
&+\cdots+z^{n-1}(f(w+(n-1) b)-z f(w+n b)) \\
& \leq(w+b)+z(w+2 b)+\cdots+z^{n-1}(w+n b)
\end{aligned}
$$

From (8) we obtain

$$
f(w+n b) \leq w+n b+a
$$

and, thus, we have for all positive integers $n$

$$
\begin{equation*}
f(w) \leq\left(1+z+\cdots+z^{n-1}+z^{n}\right) w+\left(1+2 z+\cdots+n z^{n-1}+n z^{n}\right) b+z^{n} a . \tag{11}
\end{equation*}
$$

With $w=0$ we get

$$
\begin{equation*}
a \leq\left(1+2 z+\cdots+n z^{n-1}+n z^{n}\right) b+a z^{n} . \tag{12}
\end{equation*}
$$

In view of the assumption $b<0$ we find some $n$ such that

$$
\begin{equation*}
a>(n b+a) z^{n} \tag{13}
\end{equation*}
$$

because the right hand side tends to $-\infty$ as $n \rightarrow \infty$. Now (12) and (13) give the desired contradiction and (9) is established. In addition, we have for $z=1$ the strict inequality

$$
\begin{equation*}
f(1)>0 . \tag{14}
\end{equation*}
$$

Indeed, assume that $f(1)=0$. Then setting $w=-1$ and $z=1$ in (11) leads to

$$
f(-1) \leq-(n+1)+a
$$

which is false if $n$ is sufficiently large.
To complete the proof we set $t=\min \{-a,-2 / f(1)\}$. Setting $x=1$ and $y=t$ in (7) gives

$$
\begin{equation*}
f(1-f(t)) \leq t f(1)+1 \leq-2+1=-1 . \tag{15}
\end{equation*}
$$

On the other hand, by (8) and the choice of $t$ we have $f(t) \leq t+a \leq 0$ and hence $1-f(t) \geq 1$. The inequality (9) yields

$$
f(1-f(t)) \geq 0,
$$

which contradicts (15).

## A6 USA (United States of America)

Suppose that $s_{1}, s_{2}, s_{3}, \ldots$ is a strictly increasing sequence of positive integers such that the subsequences

$$
s_{s_{1}}, s_{s_{2}}, s_{s_{3}}, \ldots \quad \text { and } \quad s_{s_{1}+1}, s_{s_{2}+1}, s_{s_{3}+1}, \ldots
$$

are both arithmetic progressions. Prove that $s_{1}, s_{2}, s_{3}, \ldots$ is itself an arithmetic progression.

Solution 1. Let $D$ be the common difference of the progression $s_{s_{1}}, s_{s_{2}}, \ldots$. Let for $n=$ $1,2, \ldots$

$$
d_{n}=s_{n+1}-s_{n} .
$$

We have to prove that $d_{n}$ is constant. First we show that the numbers $d_{n}$ are bounded. Indeed, by supposition $d_{n} \geq 1$ for all $n$. Thus, we have for all $n$

$$
d_{n}=s_{n+1}-s_{n} \leq d_{s_{n}}+d_{s_{n}+1}+\cdots+d_{s_{n+1}-1}=s_{s_{n+1}}-s_{s_{n}}=D .
$$

The boundedness implies that there exist

$$
m=\min \left\{d_{n}: n=1,2, \ldots\right\} \quad \text { and } \quad M=\max \left\{d_{n}: n=1,2, \ldots\right\}
$$

It suffices to show that $m=M$. Assume that $m<M$. Choose $n$ such that $d_{n}=m$. Considering a telescoping sum of $m=d_{n}=s_{n+1}-s_{n}$ items not greater than $M$ leads to

$$
\begin{equation*}
D=s_{s_{n+1}}-s_{s_{n}}=s_{s_{n}+m}-s_{s_{n}}=d_{s_{n}}+d_{s_{n}+1}+\cdots+d_{s_{n}+m-1} \leq m M \tag{1}
\end{equation*}
$$

and equality holds if and only if all items of the sum are equal to $M$. Now choose $n$ such that $d_{n}=M$. In the same way, considering a telescoping sum of $M$ items not less than $m$ we obtain

$$
\begin{equation*}
D=s_{s_{n+1}}-s_{s_{n}}=s_{s_{n}+M}-s_{s_{n}}=d_{s_{n}}+d_{s_{n}+1}+\cdots+d_{s_{n}+M-1} \geq M m \tag{2}
\end{equation*}
$$

and equality holds if and only if all items of the sum are equal to $m$. The inequalities (1) and (2) imply that $D=M m$ and that

$$
\begin{aligned}
d_{s_{n}}=d_{s_{n}+1}=\cdots=d_{s_{n+1}-1}=M & \text { if } d_{n}=m \\
d_{s_{n}}=d_{s_{n}+1}=\cdots=d_{s_{n+1}-1}=m & \text { if } d_{n}=M .
\end{aligned}
$$

Hence, $d_{n}=m$ implies $d_{s_{n}}=M$. Note that $s_{n} \geq s_{1}+(n-1) \geq n$ for all $n$ and moreover $s_{n}>n$ if $d_{n}=n$, because in the case $s_{n}=n$ we would have $m=d_{n}=d_{s_{n}}=M$ in contradiction to the assumption $m<M$. In the same way $d_{n}=M$ implies $d_{s_{n}}=m$ and $s_{n}>n$. Consequently, there is a strictly increasing sequence $n_{1}, n_{2}, \ldots$ such that

$$
d_{s_{n_{1}}}=M, \quad d_{s_{n_{2}}}=m, \quad d_{s_{n_{3}}}=M, \quad d_{s_{n_{4}}}=m, \quad \ldots
$$

The sequence $d_{s_{1}}, d_{s_{2}}, \ldots$ is the sequence of pairwise differences of $s_{s_{1}+1}, s_{s_{2}+1}, \ldots$ and $s_{s_{1}}, s_{s_{2}}, \ldots$, hence also an arithmetic progression. Thus $m=M$.

Solution 2. Let the integers $D$ and $E$ be the common differences of the progressions $s_{s_{1}}, s_{s_{2}}, \ldots$ and $s_{s_{1}+1}, s_{s_{2}+1}, \ldots$, respectively. Let briefly $A=s_{s_{1}}-D$ and $B=s_{s_{1}+1}-E$. Then, for all positive integers $n$,

$$
s_{s_{n}}=A+n D, \quad s_{s_{n}+1}=B+n E
$$

Since the sequence $s_{1}, s_{2}, \ldots$ is strictly increasing, we have for all positive integers $n$

$$
s_{s_{n}}<s_{s_{n}+1} \leq s_{s_{n+1}},
$$

which implies

$$
A+n D<B+n E \leq A+(n+1) D
$$

and thereby

$$
0<B-A+n(E-D) \leq D
$$

which implies $D-E=0$ and thus

$$
\begin{equation*}
0 \leq B-A \leq D \tag{3}
\end{equation*}
$$

Let $m=\min \left\{s_{n+1}-s_{n}: n=1,2, \ldots\right\}$. Then

$$
\begin{equation*}
B-A=\left(s_{s_{1}+1}-E\right)-\left(s_{s_{1}}-D\right)=s_{s_{1}+1}-s_{s_{1}} \geq m \tag{4}
\end{equation*}
$$

and

$$
\begin{equation*}
D=A+\left(s_{1}+1\right) D-\left(A+s_{1} D\right)=s_{s_{s_{1}+1}}-s_{s_{s_{1}}}=s_{B+D}-s_{A+D} \geq m(B-A) \tag{5}
\end{equation*}
$$

From (3) we consider two cases.
Case 1. $B-A=D$.
Then, for each positive integer $n, s_{s_{n}+1}=B+n D=A+(n+1) D=s_{s_{n+1}}$, hence $s_{n+1}=s_{n}+1$ and $s_{1}, s_{2}, \ldots$ is an arithmetic progression with common difference 1 .

Case 2. $B-A<D$. Choose some positive integer $N$ such that $s_{N+1}-s_{N}=m$. Then

$$
\begin{aligned}
m(A-B+D-1) & =m((A+(N+1) D)-(B+N D+1)) \\
& \leq s_{A+(N+1) D}-s_{B+N D+1}=s_{s_{s_{N+1}}}-s_{s_{s_{N}+1}+1} \\
& =\left(A+s_{N+1} D\right)-\left(B+\left(s_{N}+1\right) D\right)=\left(s_{N+1}-s_{N}\right) D+A-B-D \\
& =m D+A-B-D,
\end{aligned}
$$

i.e.,

$$
\begin{equation*}
(B-A-m)+(D-m(B-A)) \leq 0 \tag{6}
\end{equation*}
$$

The inequalities (4)-(6) imply that

$$
B-A=m \quad \text { and } \quad D=m(B-A)
$$

Assume that there is some positive integer $n$ such that $s_{n+1}>s_{n}+m$. Then $\left.m(m+1) \leq m\left(s_{n+1}-s_{n}\right) \leq s_{s_{n+1}}-s_{s_{n}}=(A+(n+1) D)-(A+n D)\right)=D=m(B-A)=m^{2}$, a contradiction. Hence $s_{1}, s_{2}, \ldots$ is an arithmetic progression with common difference $m$.

## A7 JPN (Japan)

Find all functions $f$ from the set of real numbers into the set of real numbers which satisfy for all real $x, y$ the identity

$$
f(x f(x+y))=f(y f(x))+x^{2}
$$

Solution 1. It is no hard to see that the two functions given by $f(x)=x$ and $f(x)=-x$ for all real $x$ respectively solve the functional equation. In the sequel, we prove that there are no further solutions.
Let $f$ be a function satisfying the given equation. It is clear that $f$ cannot be a constant. Let us first show that $f(0)=0$. Suppose that $f(0) \neq 0$. For any real $t$, substituting $(x, y)=\left(0, \frac{t}{f(0)}\right)$ into the given functional equation, we obtain

$$
\begin{equation*}
f(0)=f(t) \tag{1}
\end{equation*}
$$

contradicting the fact that $f$ is not a constant function. Therefore, $f(0)=0$. Next for any $t$, substituting $(x, y)=(t, 0)$ and $(x, y)=(t,-t)$ into the given equation, we get

$$
f(t f(t))=f(0)+t^{2}=t^{2}
$$

and

$$
f(t f(0))=f(-t f(t))+t^{2}
$$

respectively. Therefore, we conclude that

$$
\begin{equation*}
f(t f(t))=t^{2}, \quad f(-t f(t))=-t^{2}, \quad \text { for every real } t \tag{2}
\end{equation*}
$$

Consequently, for every real $v$, there exists a real $u$, such that $f(u)=v$. We also see that if $f(t)=0$, then $0=f(t f(t))=t^{2}$ so that $t=0$, and thus 0 is the only real number satisfying $f(t)=0$.
We next show that for any real number $s$,

$$
\begin{equation*}
f(-s)=-f(s) \tag{3}
\end{equation*}
$$

This is clear if $f(s)=0$. Suppose now $f(s)<0$, then we can find a number $t$ for which $f(s)=-t^{2}$. As $t \neq 0$ implies $f(t) \neq 0$, we can also find number $a$ such that $a f(t)=s$. Substituting $(x, y)=(t, a)$ into the given equation, we get

$$
f(t f(t+a))=f(a f(t))+t^{2}=f(s)+t^{2}=0
$$

and therefore, $t f(t+a)=0$, which implies $t+a=0$, and hence $s=-t f(t)$. Consequently, $f(-s)=f(t f(t))=t^{2}=-\left(-t^{2}\right)=-f(s)$ holds in this case.
Finally, suppose $f(s)>0$ holds. Then there exists a real number $t \neq 0$ for which $f(s)=t^{2}$. Choose a number $a$ such that $t f(a)=s$. Substituting $(x, y)=(t, a-t)$ into the given equation, we get $f(s)=f(t f(a))=f((a-t) f(t))+t^{2}=f((a-t) f(t))+f(s)$. So we have $f((a-t) f(t))=0$, from which we conclude that $(a-t) f(t)=0$. Since $f(t) \neq 0$, we get $a=t$ so that $s=t f(t)$ and thus we see $f(-s)=f(-t f(t))=-t^{2}=-f(s)$ holds in this case also. This observation finishes the proof of (3).
By substituting $(x, y)=(s, t),(x, y)=(t,-s-t)$ and $(x, y)=(-s-t, s)$ into the given equation,
we obtain

$$
\begin{array}{r}
f(s f(s+t)))=f(t f(s))+s^{2} \\
f(t f(-s))=f((-s-t) f(t))+t^{2}
\end{array}
$$

and

$$
f((-s-t) f(-t))=f(s f(-s-t))+(s+t)^{2}
$$

respectively. Using the fact that $f(-x)=-f(x)$ holds for all $x$ to rewrite the second and the third equation, and rearranging the terms, we obtain

$$
\begin{aligned}
f(t f(s))-f(s f(s+t)) & =-s^{2}, \\
f(t f(s))-f((s+t) f(t)) & =-t^{2}, \\
f((s+t) f(t))+f(s f(s+t)) & =(s+t)^{2} .
\end{aligned}
$$

Adding up these three equations now yields $2 f(t f(s))=2 t s$, and therefore, we conclude that $f(t f(s))=t s$ holds for every pair of real numbers $s, t$. By fixing $s$ so that $f(s)=1$, we obtain $f(x)=s x$. In view of the given equation, we see that $s= \pm 1$. It is easy to check that both functions $f(x)=x$ and $f(x)=-x$ satisfy the given functional equation, so these are the desired solutions.

Solution 2. As in Solution 1 we obtain (1), (2) and (3).
Now we prove that $f$ is injective. For this purpose, let us assume that $f(r)=f(s)$ for some $r \neq s$. Then, by (2)

$$
r^{2}=f(r f(r))=f(r f(s))=f((s-r) f(r))+r^{2}
$$

where the last statement follows from the given functional equation with $x=r$ and $y=s-r$. Hence, $h=(s-r) f(r)$ satisfies $f(h)=0$ which implies $h^{2}=f(h f(h))=f(0)=0$, i.e., $h=0$. Then, by $s \neq r$ we have $f(r)=0$ which implies $r=0$, and finally $f(s)=f(r)=f(0)=0$. Analogously, it follows that $s=0$ which gives the contradiction $r=s$.

To prove $|f(1)|=1$ we apply (2) with $t=1$ and also with $t=f(1)$ and obtain $f(f(1))=1$ and $(f(1))^{2}=f(f(1) \cdot f(f(1)))=f(f(1))=1$.
Now we choose $\eta \in\{-1,1\}$ with $f(1)=\eta$. Using that $f$ is odd and the given equation with $x=1, y=z$ (second equality) and with $x=-1, y=z+2$ (fourth equality) we obtain

$$
\begin{align*}
& f(z)+2 \eta=\eta(f(z \eta)+2)=\eta(f(f(z+1))+1)=\eta(-f(-f(z+1))+1) \\
& =-\eta f((z+2) f(-1))=-\eta f((z+2)(-\eta))=\eta f((z+2) \eta)=f(z+2) . \tag{4}
\end{align*}
$$

Hence,

$$
f(z+2 \eta)=\eta f(\eta z+2)=\eta(f(\eta z)+2 \eta)=f(z)+2 .
$$

Using this argument twice we obtain

$$
f(z+4 \eta)=f(z+2 \eta)+2=f(z)+4
$$

Substituting $z=2 f(x)$ we have

$$
f(2 f(x))+4=f(2 f(x)+4 \eta)=f(2 f(x+2)),
$$

where the last equality follows from (4). Applying the given functional equation we proceed to

$$
f(2 f(x+2))=f(x f(2))+4=f(2 \eta x)+4
$$

where the last equality follows again from (4) with $z=0$, i.e., $f(2)=2 \eta$. Finally, $f(2 f(x))=$ $f(2 \eta x)$ and by injectivity of $f$ we get $2 f(x)=2 \eta x$ and hence the two solutions.

## Combinatorics

## C1 NZL (New Zealand)

Consider 2009 cards, each having one gold side and one black side, lying in parallel on a long table. Initially all cards show their gold sides. Two players, standing by the same long side of the table, play a game with alternating moves. Each move consists of choosing a block of 50 consecutive cards, the leftmost of which is showing gold, and turning them all over, so those which showed gold now show black and vice versa. The last player who can make a legal move wins.
(a) Does the game necessarily end?
(b) Does there exist a winning strategy for the starting player?

Solution. (a) We interpret a card showing black as the digit 0 and a card showing gold as the digit 1. Thus each position of the 2009 cards, read from left to right, corresponds bijectively to a nonnegative integer written in binary notation of 2009 digits, where leading zeros are allowed. Each move decreases this integer, so the game must end.
(b) We show that there is no winning strategy for the starting player. We label the cards from right to left by $1, \ldots, 2009$ and consider the set $S$ of cards with labels $50 i, i=1,2, \ldots, 40$. Let $g_{n}$ be the number of cards from $S$ showing gold after $n$ moves. Obviously, $g_{0}=40$. Moreover, $\left|g_{n}-g_{n+1}\right|=1$ as long as the play goes on. Thus, after an odd number of moves, the nonstarting player finds a card from $S$ showing gold and hence can make a move. Consequently, this player always wins.

## C2 ROU (Romania)

For any integer $n \geq 2$, let $N(n)$ be the maximal number of triples $\left(a_{i}, b_{i}, c_{i}\right), i=1, \ldots, N(n)$, consisting of nonnegative integers $a_{i}, b_{i}$ and $c_{i}$ such that the following two conditions are satisfied:
(1) $a_{i}+b_{i}+c_{i}=n$ for all $i=1, \ldots, N(n)$,
(2) If $i \neq j$, then $a_{i} \neq a_{j}, b_{i} \neq b_{j}$ and $c_{i} \neq c_{j}$.

Determine $N(n)$ for all $n \geq 2$.

Comment. The original problem was formulated for $m$-tuples instead for triples. The numbers $N(m, n)$ are then defined similarly to $N(n)$ in the case $m=3$. The numbers $N(3, n)$ and $N(n, n)$ should be determined. The case $m=3$ is the same as in the present problem. The upper bound for $N(n, n)$ can be proved by a simple generalization. The construction of a set of triples attaining the bound can be easily done by induction from $n$ to $n+2$.

Solution. Let $n \geq 2$ be an integer and let $\left\{T_{1}, \ldots, T_{N}\right\}$ be any set of triples of nonnegative integers satisfying the conditions (1) and (2). Since the $a$-coordinates are pairwise distinct we have

$$
\sum_{i=1}^{N} a_{i} \geq \sum_{i=1}^{N}(i-1)=\frac{N(N-1)}{2}
$$

Analogously,

$$
\sum_{i=1}^{N} b_{i} \geq \frac{N(N-1)}{2} \quad \text { and } \quad \sum_{i=1}^{N} c_{i} \geq \frac{N(N-1)}{2}
$$

Summing these three inequalities and applying (1) yields

$$
3 \frac{N(N-1)}{2} \leq \sum_{i=1}^{N} a_{i}+\sum_{i=1}^{N} b_{i}+\sum_{i=1}^{N} c_{i}=\sum_{i=1}^{N}\left(a_{i}+b_{i}+c_{i}\right)=n N,
$$

hence $3 \frac{N-1}{2} \leq n$ and, consequently,

$$
N \leq\left\lfloor\frac{2 n}{3}\right\rfloor+1
$$

By constructing examples, we show that this upper bound can be attained, so $N(n)=\left\lfloor\frac{2 n}{3}\right\rfloor+1$.

We distinguish the cases $n=3 k-1, n=3 k$ and $n=3 k+1$ for $k \geq 1$ and present the extremal examples in form of a table.

| $n=3 k-1$ |  |  |
| :---: | :---: | :---: |
| $\left\lfloor\frac{2 n}{3}\right\rfloor+1=2 k$ |  |  |
| $a_{i}$ | $b_{i}$ | $c_{i}$ |
| 0 | $k+1$ | $2 k-2$ |
| 1 | $k+2$ | $2 k-4$ |
| $\vdots$ | $\vdots$ | $\vdots$ |
| $k-1$ | $2 k$ | 0 |
| $k$ | 0 | $2 k-1$ |
| $k+1$ | 1 | $2 k-3$ |
| $\vdots$ | $\vdots$ | $\vdots$ |
| $2 k-1$ | $k-1$ | 1 |


| $n=3 k$ |  |  |
| :---: | :---: | :---: |
| $\left\lfloor\frac{2 n}{3}\right\rfloor+1=2 k+1$ |  |  |
| $a_{i}$ | $b_{i}$ | $c_{i}$ |
| 0 | $k$ | $2 k$ |
| 1 | $k+1$ | $2 k-2$ |
| $\vdots$ | $\vdots$ | $\vdots$ |
| $k$ | $2 k$ | 0 |
| $k+1$ | 0 | $2 k-1$ |
| $k+2$ | 1 | $2 k-3$ |
| $\vdots$ | $\vdots$ | $\vdots$ |
| $2 k$ | $k-1$ | 1 |


| $n=3 k+1$ |  |  |
| :---: | :---: | :---: |
| $\left\lfloor\left\lfloor\frac{2 n}{3}\right\rfloor+1=2 k+1\right.$ |  |  |
| $a_{i}$ | $b_{i}$ | $c_{i}$ |
| 0 | $k$ | $2 k+1$ |
| 1 | $k+1$ | $2 k-1$ |
| $\vdots$ | $\vdots$ | $\vdots$ |
| $k$ | $2 k$ | 1 |
| $k+1$ | 0 | $2 k$ |
| $k+2$ | 1 | $2 k-2$ |
| $\vdots$ | $\vdots$ | $\vdots$ |
| $2 k$ | $k-1$ | 2 |

It can be easily seen that the conditions (1) and (2) are satisfied and that we indeed have $\left\lfloor\frac{2 n}{3}\right\rfloor+1$ triples in each case.

Comment. A cute combinatorial model is given by an equilateral triangle, partitioned into $n^{2}$ congruent equilateral triangles by $n-1$ equidistant parallels to each of its three sides. Two chess-like bishops placed at any two vertices of the small triangles are said to menace one another if they lie on a same parallel. The problem is to determine the largest number of bishops that can be placed so that none menaces another. A bishop may be assigned three coordinates $a, b, c$, namely the numbers of sides of small triangles they are off each of the sides of the big triangle. It is readily seen that the sum of these coordinates is always $n$, therefore fulfilling the requirements.

## C3 RUS (Russian Federation)

Let $n$ be a positive integer. Given a sequence $\varepsilon_{1}, \ldots, \varepsilon_{n-1}$ with $\varepsilon_{i}=0$ or $\varepsilon_{i}=1$ for each $i=1, \ldots, n-1$, the sequences $a_{0}, \ldots, a_{n}$ and $b_{0}, \ldots, b_{n}$ are constructed by the following rules:

$$
\begin{gathered}
a_{0}=b_{0}=1, \quad a_{1}=b_{1}=7, \\
a_{i+1}=\left\{\begin{array}{ll}
2 a_{i-1}+3 a_{i}, & \text { if } \varepsilon_{i}=0, \\
3 a_{i-1}+a_{i}, & \text { if } \varepsilon_{i}=1,
\end{array} \text { for each } i=1, \ldots, n-1,\right. \\
b_{i+1}=\left\{\begin{array}{ll}
2 b_{i-1}+3 b_{i}, & \text { if } \varepsilon_{n-i}=0, \\
3 b_{i-1}+b_{i}, & \text { if } \varepsilon_{n-i}=1,
\end{array} \text { for each } i=1, \ldots, n-1 .\right.
\end{gathered}
$$

Prove that $a_{n}=b_{n}$.

Solution. For a binary word $w=\sigma_{1} \ldots \sigma_{n}$ of length $n$ and a letter $\sigma \in\{0,1\}$ let $w \sigma=$ $\sigma_{1} \ldots \sigma_{n} \sigma$ and $\sigma w=\sigma \sigma_{1} \ldots \sigma_{n}$. Moreover let $\bar{w}=\sigma_{n} \ldots \sigma_{1}$ and let $\emptyset$ be the empty word (of length 0 and with $\bar{\emptyset}=\emptyset)$. Let $(u, v)$ be a pair of two real numbers. For binary words $w$ we define recursively the numbers $(u, v)^{w}$ as follows:

$$
\begin{gathered}
(u, v)^{\emptyset}=v, \quad(u, v)^{0}=2 u+3 v, \quad(u, v)^{1}=3 u+v, \\
(u, v)^{w \sigma \varepsilon}= \begin{cases}3(u, v)^{w}+3(u, v)^{w \sigma}, & \text { if } \varepsilon=0, \\
3(u, v)^{w}+(u, v)^{w \sigma}, & \text { if } \varepsilon=1 .\end{cases}
\end{gathered}
$$

It easily follows by induction on the length of $w$ that for all real numbers $u_{1}, v_{1}, u_{2}, v_{2}, \lambda_{1}$ and $\lambda_{2}$

$$
\begin{equation*}
\left(\lambda_{1} u_{1}+\lambda_{2} u_{2}, \lambda_{1} v_{1}+\lambda_{2} v_{2}\right)^{w}=\lambda_{1}\left(u_{1}, v_{1}\right)^{w}+\lambda_{2}\left(u_{2}, v_{2}\right)^{w} \tag{1}
\end{equation*}
$$

and that for $\varepsilon \in\{0,1\}$

$$
\begin{equation*}
(u, v)^{\varepsilon w}=\left(v,(u, v)^{\varepsilon}\right)^{w} . \tag{2}
\end{equation*}
$$

Obviously, for $n \geq 1$ and $w=\varepsilon_{1} \ldots \varepsilon_{n-1}$, we have $a_{n}=(1,7)^{w}$ and $b_{n}=(1,7)^{\bar{w}}$. Thus it is sufficient to prove that

$$
\begin{equation*}
(1,7)^{w}=(1,7)^{\bar{w}} \tag{3}
\end{equation*}
$$

for each binary word $w$. We proceed by induction on the length of $w$. The assertion is obvious if $w$ has length 0 or 1 . Now let $w \sigma \varepsilon$ be a binary word of length $n \geq 2$ and suppose that the assertion is true for all binary words of length at most $n-1$.
Note that $(2,1)^{\sigma}=7=(1,7)^{\emptyset}$ for $\sigma \in\{0,1\},(1,7)^{0}=23$, and $(1,7)^{1}=10$.
First let $\varepsilon=0$. Then in view of the induction hypothesis and the equalities (1) and (2), we obtain

$$
\begin{aligned}
&(1,7)^{w \sigma 0}=2(1,7)^{w}+3(1,7)^{w \sigma}=2(1,7)^{\bar{w}}+3(1,7)^{\sigma \bar{w}}=2(2,1)^{\sigma \bar{w}}+3(1,7)^{\sigma \bar{w}} \\
&=(7,23)^{\sigma \bar{w}}=(1,7)^{0 \sigma \bar{w}}
\end{aligned}
$$

Now let $\varepsilon=1$. Analogously, we obtain

$$
\begin{aligned}
&(1,7)^{w \sigma 1}=3(1,7)^{w}+(1,7)^{w \sigma}=3(1,7)^{\bar{w}}+(1,7)^{\sigma \bar{w}}=3(2,1)^{\sigma \bar{w}}+(1,7)^{\sigma \bar{w}} \\
&=(7,10)^{\sigma \bar{w}}=(1,7)^{1 \sigma \bar{w}}
\end{aligned}
$$

Thus the induction step is complete, (3) and hence also $a_{n}=b_{n}$ are proved.

Comment. The original solution uses the relation

$$
(1,7)^{\alpha \beta w}=\left((1,7)^{w},(1,7)^{\beta w}\right)^{\alpha}, \quad \alpha, \beta \in\{0,1\},
$$

which can be proved by induction on the length of $w$. Then (3) also follows by induction on the length of $w$ :

$$
(1,7)^{\alpha \beta w}=\left((1,7)^{w},(1,7)^{\beta w}\right)^{\alpha}=\left((1,7)^{\bar{w}},(1,7)^{\bar{w} \beta}\right)^{\alpha}=(1,7)^{\bar{w} \beta \alpha} .
$$

Here $w$ may be the empty word.

## C4 NLD (Netherlands)

For an integer $m \geq 1$, we consider partitions of a $2^{m} \times 2^{m}$ chessboard into rectangles consisting of cells of the chessboard, in which each of the $2^{m}$ cells along one diagonal forms a separate rectangle of side length 1 . Determine the smallest possible sum of rectangle perimeters in such a partition.

Solution 1. For a $k \times k$ chessboard, we introduce in a standard way coordinates of the vertices of the cells and assume that the cell $C_{i j}$ in row $i$ and column $j$ has vertices $(i-1, j-1),(i-$ $1, j),(i, j-1),(i, j)$, where $i, j \in\{1, \ldots, k\}$. Without loss of generality assume that the cells $C_{i i}$, $i=1, \ldots, k$, form a separate rectangle. Then we may consider the boards $B_{k}=\bigcup_{1 \leq i<j \leq k} C_{i j}$ below that diagonal and the congruent board $B_{k}^{\prime}=\bigcup_{1 \leq j<i \leq k} C_{i j}$ above that diagonal separately because no rectangle can simultaneously cover cells from $B_{k}$ and $B_{k}^{\prime}$. We will show that for $k=2^{m}$ the smallest total perimeter of a rectangular partition of $B_{k}$ is $m 2^{m+1}$. Then the overall answer to the problem is $2 \cdot m 2^{m+1}+4 \cdot 2^{m}=(m+1) 2^{m+2}$.
First we inductively construct for $m \geq 1$ a partition of $B_{2^{m}}$ with total perimeter $m 2^{m+1}$. If $m=0$, the board $B_{2^{m}}$ is empty and the total perimeter is 0 . For $m \geq 0$, the board $B_{2^{m+1}}$ consists of a $2^{m} \times 2^{m}$ square in the lower right corner with vertices $\left(2^{m}, 2^{m}\right),\left(2^{m}, 2^{m+1}\right),\left(2^{m+1}, 2^{m}\right)$, $\left(2^{m+1}, 2^{m+1}\right)$ to which two boards congruent to $B_{2^{m}}$ are glued along the left and the upper margin. The square together with the inductive partitions of these two boards yield a partition with total perimeter $4 \cdot 2^{m}+2 \cdot m 2^{m+1}=(m+1) 2^{m+2}$ and the induction step is complete.
Let

$$
D_{k}=2 k \log _{2} k
$$

Note that $D_{k}=m 2^{m+1}$ if $k=2^{m}$. Now we show by induction on $k$ that the total perimeter of a rectangular partition of $B_{k}$ is at least $D_{k}$. The case $k=1$ is trivial (see $m=0$ from above). Let the assertion be true for all positive integers less than $k$. We investigate a fixed rectangular partition of $B_{k}$ that attains the minimal total perimeter. Let $R$ be the rectangle that covers the cell $C_{1 k}$ in the lower right corner. Let $(i, j)$ be the upper left corner of $R$. First we show that $i=j$. Assume that $i<j$. Then the line from $(i, j)$ to $(i+1, j)$ or from $(i, j)$ to $(i, j-1)$ must belong to the boundary of some rectangle in the partition. Without loss of generality assume that this is the case for the line from $(i, j)$ to $(i+1, j)$.
Case 1. No line from $(i, l)$ to $(i+1, l)$ where $j<l<k$ belongs to the boundary of some rectangle of the partition.
Then there is some rectangle $R^{\prime}$ of the partition that has with $R$ the common side from $(i, j)$ to $(i, k)$. If we join these two rectangles to one rectangle we get a partition with smaller total perimeter, a contradiction.
Case 2. There is some $l$ such that $j<l<k$ and the line from $(i, l)$ to $(i+1, l)$ belongs to the boundary of some rectangle of the partition.
Then we replace the upper side of $R$ by the line $(i+1, j)$ to $(i+1, k)$ and for the rectangles whose lower side belongs to the line from $(i, j)$ to $(i, k)$ we shift the lower side upwards so that the new lower side belongs to the line from $(i+1, j)$ to $(i+1, k)$. In such a way we obtain a rectangular partition of $B_{k}$ with smaller total perimeter, a contradiction.
Now the fact that the upper left corner of $R$ has the coordinates $(i, i)$ is established. Consequently, the partition consists of $R$, of rectangles of a partition of a board congruent to $B_{i}$ and of rectangles of a partition of a board congruent to $B_{k-i}$. By the induction hypothesis, its total
perimeter is at least

$$
\begin{equation*}
2(k-i)+2 i+D_{i}+D_{k-i} \geq 2 k+2 i \log _{2} i+2(k-i) \log _{2}(k-i) . \tag{1}
\end{equation*}
$$

Since the function $f(x)=2 x \log _{2} x$ is convex for $x>0$, Jensen's inequality immediately shows that the minimum of the right hand sight of (1) is attained for $i=k / 2$. Hence the total perimeter of the optimal partition of $B_{k}$ is at least $2 k+2 k / 2 \log _{2} k / 2+2(k / 2) \log _{2}(k / 2)=D_{k}$.

Solution 2. We start as in Solution 1 and present another proof that $m 2^{m+1}$ is a lower bound for the total perimeter of a partition of $B_{2^{m}}$ into $n$ rectangles. Let briefly $M=2^{m}$. For $1 \leq i \leq M$, let $r_{i}$ denote the number of rectangles in the partition that cover some cell from row $i$ and let $c_{j}$ be the number of rectangles that cover some cell from column $j$. Note that the total perimeter $p$ of all rectangles in the partition is

$$
p=2\left(\sum_{i=1}^{M} r_{i}+\sum_{i=1}^{M} c_{i}\right) .
$$

No rectangle can simultaneously cover cells from row $i$ and from column $i$ since otherwise it would also cover the cell $C_{i i}$. We classify subsets $S$ of rectangles of the partition as follows. We say that $S$ is of type $i, 1 \leq i \leq M$, if $S$ contains all $r_{i}$ rectangles that cover some cell from row $i$, but none of the $c_{i}$ rectangles that cover some cell from column $i$. Altogether there are $2^{n-r_{i}-c_{i}}$ subsets of type $i$. Now we show that no subset $S$ can be simultaneously of type $i$ and of type $j$ if $i \neq j$. Assume the contrary and let without loss of generality $i<j$. The cell $C_{i j}$ must be covered by some rectangle $R$. The subset $S$ is of type $i$, hence $R$ is contained in $S$. $S$ is of type $j$, thus $R$ does not belong to $S$, a contradiction. Since there are $2^{n}$ subsets of rectangles of the partition, we infer

$$
\begin{equation*}
2^{n} \geq \sum_{i=1}^{M} 2^{n-r_{i}-c_{i}}=2^{n} \sum_{i=1}^{M} 2^{-\left(r_{i}+c_{i}\right)} . \tag{2}
\end{equation*}
$$

By applying Jensen's inequality to the convex function $f(x)=2^{-x}$ we derive

$$
\begin{equation*}
\frac{1}{M} \sum_{i=1}^{M} 2^{-\left(r_{i}+c_{i}\right)} \geq 2^{-\frac{1}{M} \sum_{i=1}^{M}\left(r_{i}+c_{i}\right)}=2^{-\frac{p}{2 M}} \tag{3}
\end{equation*}
$$

From (2) and (3) we obtain

$$
1 \geq M 2^{-\frac{p}{2 M}}
$$

and equivalently

$$
p \geq m 2^{m+1}
$$

## C5 NLD (Netherlands)

Five identical empty buckets of 2-liter capacity stand at the vertices of a regular pentagon. Cinderella and her wicked Stepmother go through a sequence of rounds: At the beginning of every round, the Stepmother takes one liter of water from the nearby river and distributes it arbitrarily over the five buckets. Then Cinderella chooses a pair of neighboring buckets, empties them into the river, and puts them back. Then the next round begins. The Stepmother's goal is to make one of these buckets overflow. Cinderella's goal is to prevent this. Can the wicked Stepmother enforce a bucket overflow?

Solution 1. No, the Stepmother cannot enforce a bucket overflow and Cinderella can keep playing forever. Throughout we denote the five buckets by $B_{0}, B_{1}, B_{2}, B_{3}$, and $B_{4}$, where $B_{k}$ is adjacent to bucket $B_{k-1}$ and $B_{k+1}(k=0,1,2,3,4)$ and all indices are taken modulo 5 . Cinderella enforces that the following three conditions are satisfied at the beginning of every round:
(1) Two adjacent buckets (say $B_{1}$ and $B_{2}$ ) are empty.
(2) The two buckets standing next to these adjacent buckets (here $B_{0}$ and $B_{3}$ ) have total contents at most 1.
(3) The remaining bucket (here $B_{4}$ ) has contents at most 1 .

These conditions clearly hold at the beginning of the first round, when all buckets are empty.
Assume that Cinderella manages to maintain them until the beginning of the $r$-th round $(r \geq 1)$. Denote by $x_{k}(k=0,1,2,3,4)$ the contents of bucket $B_{k}$ at the beginning of this round and by $y_{k}$ the corresponding contents after the Stepmother has distributed her liter of water in this round.
By the conditions, we can assume $x_{1}=x_{2}=0, x_{0}+x_{3} \leq 1$ and $x_{4} \leq 1$. Then, since the Stepmother adds one liter, we conclude $y_{0}+y_{1}+y_{2}+y_{3} \leq 2$. This inequality implies $y_{0}+y_{2} \leq 1$ or $y_{1}+y_{3} \leq 1$. For reasons of symmetry, we only consider the second case.
Then Cinderella empties buckets $B_{0}$ and $B_{4}$.
At the beginning of the next round $B_{0}$ and $B_{4}$ are empty (condition (1) is fulfilled), due to $y_{1}+y_{3} \leq 1$ condition (2) is fulfilled and finally since $x_{2}=0$ we also must have $y_{2} \leq 1$ (condition (3) is fulfilled).

Therefore, Cinderella can indeed manage to maintain the three conditions (1)-(3) also at the beginning of the $(r+1)$-th round. By induction, she thus manages to maintain them at the beginning of every round. In particular she manages to keep the contents of every single bucket at most 1 liter. Therefore, the buckets of 2-liter capacity will never overflow.

Solution 2. We prove that Cinderella can maintain the following two conditions and hence she can prevent the buckets from overflow:
(1') Every two non-adjacent buckets contain a total of at most 1.
(2') The total contents of all five buckets is at most $\frac{3}{2}$.
We use the same notations as in the first solution. The two conditions again clearly hold at the beginning. Assume that Cinderella maintained these two conditions until the beginning of the $r$-th round. A pair of non-neighboring buckets $\left(B_{i}, B_{i+2}\right), i=0,1,2,3,4$ is called critical
if $y_{i}+y_{i+2}>1$. By condition ( $2^{\prime}$ ), after the Stepmother has distributed her water we have $y_{0}+y_{1}+y_{2}+y_{3}+y_{4} \leq \frac{5}{2}$. Therefore,

$$
\left(y_{0}+y_{2}\right)+\left(y_{1}+y_{3}\right)+\left(y_{2}+y_{4}\right)+\left(y_{3}+y_{0}\right)+\left(y_{4}+y_{1}\right)=2\left(y_{0}+y_{1}+y_{2}+y_{3}+y_{4}\right) \leq 5
$$

and hence there is a pair of non-neighboring buckets which is not critical, say $\left(B_{0}, B_{2}\right)$. Now, if both of the pairs $\left(B_{3}, B_{0}\right)$ and $\left(B_{2}, B_{4}\right)$ are critical, we must have $y_{1}<\frac{1}{2}$ and Cinderella can empty the buckets $B_{3}$ and $B_{4}$. This clearly leaves no critical pair of buckets and the total contents of all the buckets is then $y_{1}+\left(y_{0}+y_{2}\right) \leq \frac{3}{2}$. Therefore, conditions ( $1^{\prime}$ ) and (2') are fulfilled.

Now suppose that without loss of generality the pair $\left(B_{3}, B_{0}\right)$ is not critical. If in this case $y_{0} \leq \frac{1}{2}$, then one of the inequalities $y_{0}+y_{1}+y_{2} \leq \frac{3}{2}$ and $y_{0}+y_{3}+y_{4} \leq \frac{3}{2}$ must hold. But then Cinderella can empty $B_{3}$ and $B_{4}$ or $B_{1}$ and $B_{2}$, respectively and clearly fulfill the conditions.
Finally consider the case $y_{0}>\frac{1}{2}$. By $y_{0}+y_{1}+y_{2}+y_{3}+y_{4} \leq \frac{5}{2}$, at least one of the pairs $\left(B_{1}, B_{3}\right)$ and $\left(B_{2}, B_{4}\right)$ is not critical. Without loss of generality let this be the pair $\left(B_{1}, B_{3}\right)$. Since the pair $\left(B_{3}, B_{0}\right)$ is not critical and $y_{0}>\frac{1}{2}$, we must have $y_{3} \leq \frac{1}{2}$. But then, as before, Cinderella can maintain the two conditions at the beginning of the next round by either emptying $B_{1}$ and $B_{2}$ or $B_{4}$ and $B_{0}$.

Comments on GREEDY approaches. A natural approach for Cinderella would be a GREEDY strategy as for example: Always remove as much water as possible from the system. It is straightforward to prove that GREEDY can avoid buckets of capacity $\frac{5}{2}$ from overflowing: If before the Stepmothers move one has $x_{0}+x_{1}+x_{2}+x_{3}+x_{4} \leq \frac{3}{2}$ then after her move the inequality $Y=y_{0}+y_{1}+y_{2}+y_{3}+y_{4} \leq \frac{5}{2}$ holds. If now Cinderella removes the two adjacent buckets with maximum total contents she removes at least $\frac{2 Y}{5}$ and thus the remaining buckets contain at most $\frac{3}{5} \cdot Y \leq \frac{3}{2}$.
But GREEDY is in general not strong enough to settle this problem as can be seen in the following example:

- In an initial phase, the Stepmother brings all the buckets (after her move) to contents of at least $\frac{1}{2}-2 \epsilon$, where $\epsilon$ is an arbitrary small positive number. This can be done by always splitting the 1 liter she has to distribute so that all buckets have the same contents. After her $r$-th move the total contents of each of the buckets is then $c_{r}$ with $c_{1}=1$ and $c_{r+1}=1+\frac{3}{5} \cdot c_{r}$ and hence $c_{r}=\frac{5}{2}-\frac{3}{2} \cdot\left(\frac{3}{5}\right)^{r-1}$. So the contents of each single bucket indeed approaches $\frac{1}{2}$ (from below). In particular, any two adjacent buckets have total contents strictly less than 1 which enables the Stepmother to always refill the buckets that Cinderella just emptied and then distribute the remaining water evenly over all buckets.
- After that phase GREEDY faces a situation like this ( $\frac{1}{2}-2 \epsilon, \frac{1}{2}-2 \epsilon, \frac{1}{2}-2 \epsilon, \frac{1}{2}-2 \epsilon, \frac{1}{2}-2 \epsilon$ ) and leaves a situation of the form $\left(x_{0}, x_{1}, x_{2}, x_{3}, x_{4}\right)=\left(\frac{1}{2}-2 \epsilon, \frac{1}{2}-2 \epsilon, \frac{1}{2}-2 \epsilon, 0,0\right)$.
- Then the Stepmother can add the amounts $\left(0, \frac{1}{4}+\epsilon, \epsilon, \frac{3}{4}-2 \epsilon, 0\right)$ to achieve a situation like this: $\left(y_{0}, y_{1}, y_{2}, y_{3}, y_{4}\right)=\left(\frac{1}{2}-2 \epsilon, \frac{3}{4}-\epsilon, \frac{1}{2}-\epsilon, \frac{3}{4}-2 \epsilon, 0\right)$.
- Now $B_{1}$ and $B_{2}$ are the adjacent buckets with the maximum total contents and thus GREEDY empties them to yield $\left(x_{0}, x_{1}, x_{2}, x_{3}, x_{4}\right)=\left(\frac{1}{2}-2 \epsilon, 0,0, \frac{3}{4}-2 \epsilon, 0\right)$.
- Then the Stepmother adds $\left(\frac{5}{8}, 0,0, \frac{3}{8}, 0\right)$, which yields $\left(\frac{9}{8}-2 \epsilon, 0,0, \frac{9}{8}-2 \epsilon, 0\right)$.
- Now GREEDY can only empty one of the two nonempty buckets and in the next step the Stepmother adds her liter to the other bucket and brings it to $\frac{17}{8}-2 \epsilon$, i.e. an overflow.

A harder variant. Five identical empty buckets of capacity $b$ stand at the vertices of a regular pentagon. Cinderella and her wicked Stepmother go through a sequence of rounds: At the beginning of every round, the Stepmother takes one liter of water from the nearby river and distributes it arbitrarily over the five buckets. Then Cinderella chooses a pair of neighboring buckets, empties them into the river, and puts them back. Then the next round begins. The Stepmother's goal is to make one of these buckets overflow. Cinderella's goal is to prevent this. Determine all bucket capacities $b$ for which the Stepmother can enforce a bucket to overflow.

Solution to the harder variant. The answer is $b<2$.
The previous proof shows that for all $b \geq 2$ the Stepmother cannot enforce overflowing. Now if $b<2$, let $R$ be a positive integer such that $b<2-2^{1-R}$. In the first $R$ rounds the Stepmother now ensures that at least one of the (nonadjacent) buckets $B_{1}$ and $B_{3}$ have contents of at least $1-2^{1-r}$ at the beginning of round $r(r=1,2, \ldots, R)$. This is trivial for $r=1$ and if it holds at the beginning of round $r$, she can fill the bucket which contains at least $1-2^{1-r}$ liters with another $2^{-r}$ liters and put the rest of her water - $1-2^{-r}$ liters - in the other bucket. As Cinderella now can remove the water of at most one of the two buckets, the other bucket carries its contents into the next round.

At the beginning of the $R$-th round there are $1-2^{1-R}$ liters in $B_{1}$ or $B_{3}$. The Stepmother puts the entire liter into that bucket and produces an overflow since $b<2-2^{1-R}$.

## C6 BGR (Bulgaria)

On a $999 \times 999$ board a limp rook can move in the following way: From any square it can move to any of its adjacent squares, i.e. a square having a common side with it, and every move must be a turn, i.e. the directions of any two consecutive moves must be perpendicular. A nonintersecting route of the limp rook consists of a sequence of pairwise different squares that the limp rook can visit in that order by an admissible sequence of moves. Such a non-intersecting route is called cyclic, if the limp rook can, after reaching the last square of the route, move directly to the first square of the route and start over.
How many squares does the longest possible cyclic, non-intersecting route of a limp rook visit?

Solution. The answer is $998^{2}-4=4 \cdot\left(499^{2}-1\right)$ squares.
First we show that this number is an upper bound for the number of cells a limp rook can visit. To do this we color the cells with four colors $A, B, C$ and $D$ in the following way: for $(i, j) \equiv(0,0) \bmod 2$ use $A$, for $(i, j) \equiv(0,1) \bmod 2$ use $B$, for $(i, j) \equiv(1,0) \bmod 2$ use $C$ and for $(i, j) \equiv(1,1) \bmod 2$ use $D$. From an $A$-cell the rook has to move to a $B$-cell or a $C$-cell. In the first case, the order of the colors of the cells visited is given by $A, B, D, C, A, B, D, C, A, \ldots$, in the second case it is $A, C, D, B, A, C, D, B, A, \ldots$ Since the route is closed it must contain the same number of cells of each color. There are only $499^{2} A$-cells. In the following we will show that the rook cannot visit all the $A$-cells on its route and hence the maximum possible number of cells in a route is $4 \cdot\left(499^{2}-1\right)$.
Assume that the route passes through every single $A$-cell. Color the $A$-cells in black and white in a chessboard manner, i.e. color any two $A$-cells at distance 2 in different color. Since the number of $A$-cells is odd the rook cannot always alternate between visiting black and white $A$-cells along its route. Hence there are two $A$-cells of the same color which are four rook-steps apart that are visited directly one after the other. Let these two $A$-cells have row and column numbers $(a, b)$ and $(a+2, b+2)$ respectively.


There is up to reflection only one way the rook can take from $(a, b)$ to $(a+2, b+2)$. Let this way be $(a, b) \rightarrow(a, b+1) \rightarrow(a+1, b+1) \rightarrow(a+1, b+2) \rightarrow(a+2, b+2)$. Also let without loss of generality the color of the cell $(a, b+1)$ be $B$ (otherwise change the roles of columns and rows).
Now consider the $A$-cell $(a, b+2)$. The only way the rook can pass through it is via $(a-1, b+2) \rightarrow$ $(a, b+2) \rightarrow(a, b+3)$ in this order, since according to our assumption after every $A$-cell the rook passes through a $B$-cell. Hence, to connect these two parts of the path, there must be
a path connecting the cell $(a, b+3)$ and $(a, b)$ and also a path connecting $(a+2, b+2)$ and $(a-1, b+2)$.

But these four cells are opposite vertices of a convex quadrilateral and the paths are outside of that quadrilateral and hence they must intersect. This is due to the following fact:

The path from $(a, b)$ to $(a, b+3)$ together with the line segment joining these two cells form a closed loop that has one of the cells $(a-1, b+2)$ and $(a+2, b+2)$ in its inside and the other one on the outside. Thus the path between these two points must cross the previous path.
But an intersection is only possible if a cell is visited twice. This is a contradiction.
Hence the number of cells visited is at most $4 \cdot\left(499^{2}-1\right)$.
The following picture indicates a recursive construction for all $n \times n$-chessboards with $n \equiv 3$ $\bmod 4$ which clearly yields a path that misses exactly one $A$-cell (marked with a dot, the center cell of the $15 \times 15$-chessboard) and hence, in the case of $n=999$ crosses exactly $4 \cdot\left(499^{2}-1\right)$ cells.


Combinatorics

## C7 RUS (Russian Federation)

Variant 1. A grasshopper jumps along the real axis. He starts at point 0 and makes 2009 jumps to the right with lengths $1,2, \ldots, 2009$ in an arbitrary order. Let $M$ be a set of 2008 positive integers less than $1005 \cdot 2009$. Prove that the grasshopper can arrange his jumps in such a way that he never lands on a point from $M$.

Variant 2. Let $n$ be a nonnegative integer. A grasshopper jumps along the real axis. He starts at point 0 and makes $n+1$ jumps to the right with pairwise different positive integral lengths $a_{1}, a_{2}, \ldots, a_{n+1}$ in an arbitrary order. Let $M$ be a set of $n$ positive integers in the interval $(0, s)$, where $s=a_{1}+a_{2}+\cdots+a_{n+1}$. Prove that the grasshopper can arrange his jumps in such a way that he never lands on a point from $M$.

Solution of Variant 1. We construct the set of landing points of the grasshopper.
Case 1. $M$ does not contain numbers divisible by 2009.
We fix the numbers $2009 k$ as landing points, $k=1,2, \ldots, 1005$. Consider the open intervals $I_{k}=(2009(k-1), 2009 k), k=1,2, \ldots, 1005$. We show that we can choose exactly one point outside of $M$ as a landing point in 1004 of these intervals such that all lengths from 1 to 2009 are realized. Since there remains one interval without a chosen point, the length 2009 indeed will appear. Each interval has length 2009, hence a new landing point in an interval yields with a length $d$ also the length $2009-d$. Thus it is enough to implement only the lengths from $D=\{1,2, \ldots, 1004\}$. We will do this in a greedy way. Let $n_{k}, k=1,2, \ldots, 1005$, be the number of elements of $M$ that belong to the interval $I_{k}$. We order these numbers in a decreasing way, so let $p_{1}, p_{2}, \ldots, p_{1005}$ be a permutation of $\{1,2, \ldots, 1005\}$ such that $n_{p_{1}} \geq n_{p_{2}} \geq \cdots \geq n_{p_{1005}}$. In $I_{p_{1}}$ we do not choose a landing point. Assume that landing points have already been chosen in the intervals $I_{p_{2}}, \ldots, I_{p_{m}}$ and the lengths $d_{2}, \ldots, d_{m}$ from $D$ are realized, $m=1, \ldots, 1004$. We show that there is some $d \in D \backslash\left\{d_{2}, \ldots, d_{m}\right\}$ that can be implemented with a new landing point in $I_{p_{m+1}}$. Assume the contrary. Then the $1004-(m-1)$ other lengths are obstructed by the $n_{p_{m+1}}$ points of $M$ in $I_{p_{m+1}}$. Each length $d$ can be realized by two landing points, namely $2009\left(p_{m+1}-1\right)+d$ and $2009 p_{m+1}-d$, hence

$$
\begin{equation*}
n_{p_{m+1}} \geq 2(1005-m) \tag{1}
\end{equation*}
$$

Moreover, since $|M|=2008=n_{1}+\cdots+n_{1005}$,

$$
\begin{equation*}
2008 \geq n_{p_{1}}+n_{p_{2}}+\cdots+n_{p_{m+1}} \geq(m+1) n_{p_{m+1}} . \tag{2}
\end{equation*}
$$

Consequently, by (1) and (2),

$$
2008 \geq 2(m+1)(1005-m) .
$$

The right hand side of the last inequality obviously attains its minimum for $m=1004$ and this minimum value is greater than 2008, a contradiction.
Case 2. $M$ does contain a number $\mu$ divisible by 2009.
By the pigeonhole principle there exists some $r \in\{1, \ldots, 2008\}$ such that $M$ does not contain numbers with remainder $r$ modulo 2009. We fix the numbers $2009(k-1)+r$ as landing points, $k=1,2, \ldots, 1005$. Moreover, $1005 \cdot 2009$ is a landing point. Consider the open intervals
$I_{k}=(2009(k-1)+r, 2009 k+r), k=1,2, \ldots, 1004$. Analogously to Case 1 , it is enough to show that we can choose in 1003 of these intervals exactly one landing point outside of $M \backslash\{\mu\}$ such that each of the lengths of $D=\{1,2, \ldots, 1004\} \backslash\{r\}$ are implemented. Note that $r$ and $2009-r$ are realized by the first and last jump and that choosing $\mu$ would realize these two differences again. Let $n_{k}, k=1,2, \ldots, 1004$, be the number of elements of $M \backslash\{\mu\}$ that belong to the interval $I_{k}$ and $p_{1}, p_{2}, \ldots, p_{1004}$ be a permutation of $\{1,2, \ldots, 1004\}$ such that $n_{p_{1}} \geq n_{p_{2}} \geq \cdots \geq n_{p_{1004}}$. With the same reasoning as in Case 1 we can verify that a greedy choice of the landing points in $I_{p_{2}}, I_{p_{3}}, \ldots, I_{p_{1004}}$ is possible. We only have to replace (1) by

$$
n_{p_{m+1}} \geq 2(1004-m)
$$

( $D$ has one element less) and (2) by

$$
2007 \geq n_{p_{1}}+n_{p_{2}}+\cdots+n_{p_{m+1}} \geq(m+1) n_{p_{m+1}}
$$

Comment. The cardinality 2008 of $M$ in the problem is the maximum possible value. For $M=\{1,2, \ldots, 2009\}$, the grasshopper necessarily lands on a point from $M$.

Solution of Variant 2. First of all we remark that the statement in the problem implies a strengthening of itself: Instead of $|M|=n$ it is sufficient to suppose that $|M \cap(0, s-\bar{a}]| \leq n$, where $\bar{a}=\min \left\{a_{1}, a_{2}, \ldots, a_{n+1}\right\}$. This fact will be used in the proof.
We prove the statement by induction on $n$. The case $n=0$ is obvious. Let $n>0$ and let the assertion be true for all nonnegative integers less than $n$. Moreover let $a_{1}, a_{2}, \ldots, a_{n+1}, s$ and $M$ be given as in the problem. Without loss of generality we may assume that $a_{n+1}<a_{n}<$ $\cdots<a_{2}<a_{1}$. Set

$$
T_{k}=\sum_{i=1}^{k} a_{i} \quad \text { for } k=0,1, \ldots, n+1
$$

Note that $0=T_{0}<T_{1}<\cdots<T_{n+1}=s$. We will make use of the induction hypothesis as follows:

Claim 1. It suffices to show that for some $m \in\{1,2, \ldots, n+1\}$ the grasshopper is able to do at least $m$ jumps without landing on a point of $M$ and, in addition, after these $m$ jumps he has jumped over at least $m$ points of $M$.
Proof. Note that $m=n+1$ is impossible by $|M|=n$. Now set $n^{\prime}=n-m$. Then $0 \leq n^{\prime}<n$. The remaining $n^{\prime}+1$ jumps can be carried out without landing on one of the remaining at most $n^{\prime}$ forbidden points by the induction hypothesis together with a shift of the origin. This proves the claim.
An integer $k \in\{1,2, \ldots, n+1\}$ is called smooth, if the grasshopper is able to do $k$ jumps with the lengths $a_{1}, a_{2}, \ldots, a_{k}$ in such a way that he never lands on a point of $M$ except for the very last jump, when he may land on a point of $M$.
Obviously, 1 is smooth. Thus there is a largest number $k^{*}$, such that all the numbers $1,2, \ldots, k^{*}$ are smooth. If $k^{*}=n+1$, the proof is complete. In the following let $k^{*} \leq n$.
Claim 2. We have

$$
\begin{equation*}
T_{k^{*}} \in M \quad \text { and } \quad\left|M \cap\left(0, T_{k^{*}}\right)\right| \geq k^{*} . \tag{3}
\end{equation*}
$$

Proof. In the case $T_{k^{*}} \notin M$ any sequence of jumps that verifies the smoothness of $k^{*}$ can be extended by appending $a_{k^{*}+1}$, which is a contradiction to the maximality of $k^{*}$. Therefore we have $T_{k^{*}} \in M$. If $\left|M \cap\left(0, T_{k^{*}}\right)\right|<k^{*}$, there exists an $l \in\left\{1,2, \ldots, k^{*}\right\}$ with $T_{k^{*}+1}-a_{l} \notin M$. By the induction hypothesis with $k^{*}-1$ instead of $n$, the grasshopper is able to reach $T_{k^{*}+1}-a_{l}$
with $k^{*}$ jumps of lengths from $\left\{a_{1}, a_{2}, \ldots, a_{k^{*}+1}\right\} \backslash\left\{a_{l}\right\}$ without landing on any point of $M$. Therefore $k^{*}+1$ is also smooth, which is a contradiction to the maximality of $k^{*}$. Thus Claim 2 is proved.
Now, by Claim 2, there exists a smallest integer $\bar{k} \in\left\{1,2, \ldots, k^{*}\right\}$ with

$$
T_{\bar{k}} \in M \quad \text { and } \quad\left|M \cap\left(0, T_{\bar{k}}\right)\right| \geq \bar{k} .
$$

Claim 3. It is sufficient to consider the case

$$
\begin{equation*}
\left|M \cap\left(0, T_{\bar{k}-1}\right]\right| \leq \bar{k}-1 . \tag{4}
\end{equation*}
$$

Proof. If $\bar{k}=1$, then (4) is clearly satisfied. In the following let $\bar{k}>1$. If $T_{\bar{k}-1} \in M$, then (4) follows immediately by the minimality of $\bar{k}$. If $T_{\bar{k}-1} \notin M$, by the smoothness of $\bar{k}-1$, we obtain a situation as in Claim 1 with $m=\bar{k}-1$ provided that $\left|M \cap\left(0, T_{\bar{k}-1}\right]\right| \geq \bar{k}-1$. Hence, we may even restrict ourselves to $\left|M \cap\left(0, T_{\bar{k}-1}\right]\right| \leq \bar{k}-2$ in this case and Claim 3 is proved.
Choose an integer $v \geq 0$ with $\left|M \cap\left(0, T_{\bar{k}}\right)\right|=\bar{k}+v$. Let $r_{1}>r_{2}>\cdots>r_{l}$ be exactly those indices $r$ from $\{\bar{k}+1, \bar{k}+2, \ldots, n+1\}$ for which $T_{\bar{k}}+a_{r} \notin M$. Then

$$
n=|M|=\left|M \cap\left(0, T_{\bar{k}}\right)\right|+1+\left|M \cap\left(T_{\bar{k}}, s\right)\right| \geq \bar{k}+v+1+(n+1-\bar{k}-l)
$$

and consequently $l \geq v+2$. Note that
$T_{\bar{k}}+a_{r_{1}}-a_{1}<T_{\bar{k}}+a_{r_{1}}-a_{2}<\cdots<T_{\bar{k}}+a_{r_{1}}-a_{\bar{k}}<T_{\bar{k}}+a_{r_{2}}-a_{\bar{k}}<\cdots<T_{\bar{k}}+a_{r_{v+2}}-a_{\bar{k}}<T_{\bar{k}}$ and that this are $\bar{k}+v+1$ numbers from $\left(0, T_{\bar{k}}\right)$. Therefore we find some $r \in\{\bar{k}+1, \bar{k}+$ $2, \ldots, n+1\}$ and some $s \in\{1,2, \ldots, \bar{k}\}$ with $T_{\bar{k}}+a_{r} \notin M$ and $T_{\bar{k}}+a_{r}-a_{s} \notin M$. Consider the set of jump lengths $B=\left\{a_{1}, a_{2}, \ldots, a_{\bar{k}}, a_{r}\right\} \backslash\left\{a_{s}\right\}$. We have

$$
\sum_{x \in B} x=T_{\bar{k}}+a_{r}-a_{s}
$$

and

$$
T_{\bar{k}}+a_{r}-a_{s}-\min (B)=T_{\bar{k}}-a_{s} \leq T_{\bar{k}-1} .
$$

By (4) and the strengthening, mentioned at the very beginning with $\bar{k}-1$ instead of $n$, the grasshopper is able to reach $T_{\bar{k}}+a_{r}-a_{s}$ by $\bar{k}$ jumps with lengths from $B$ without landing on any point of $M$. From there he is able to jump to $T_{\bar{k}}+a_{r}$ and therefore we reach a situation as in Claim 1 with $m=\bar{k}+1$, which completes the proof.

## C8 AUT (Austria)

For any integer $n \geq 2$, we compute the integer $h(n)$ by applying the following procedure to its decimal representation. Let $r$ be the rightmost digit of $n$.
(1) If $r=0$, then the decimal representation of $h(n)$ results from the decimal representation of $n$ by removing this rightmost digit 0 .
(2) If $1 \leq r \leq 9$ we split the decimal representation of $n$ into a maximal right part $R$ that solely consists of digits not less than $r$ and into a left part $L$ that either is empty or ends with a digit strictly smaller than $r$. Then the decimal representation of $h(n)$ consists of the decimal representation of $L$, followed by two copies of the decimal representation of $R-1$. For instance, for the number $n=17,151,345,543$, we will have $L=17,151, R=345,543$ and $h(n)=17,151,345,542,345,542$.
Prove that, starting with an arbitrary integer $n \geq 2$, iterated application of $h$ produces the integer 1 after finitely many steps.

Solution 1. We identify integers $n \geq 2$ with the digit-strings, briefly strings, of their decimal representation and extend the definition of $h$ to all non-empty strings with digits from 0 to 9. We recursively define ten functions $f_{0}, \ldots, f_{9}$ that map some strings into integers for $k=$ $9,8, \ldots, 1,0$. The function $f_{9}$ is only defined on strings $x$ (including the empty string $\varepsilon$ ) that entirely consist of nines. If $x$ consists of $m$ nines, then $f_{9}(x)=m+1, m=0,1, \ldots$. For $k \leq 8$, the domain of $f_{k}(x)$ is the set of all strings consisting only of digits that are $\geq k$. We write $x$ in the form $x_{0} k x_{1} k x_{2} k \ldots x_{m-1} k x_{m}$ where the strings $x_{s}$ only consist of digits $\geq k+1$. Note that some of these strings might equal the empty string $\varepsilon$ and that $m=0$ is possible, i.e. the digit $k$ does not appear in $x$. Then we define

$$
f_{k}(x)=\sum_{s=0}^{m} 4^{f_{k+1}\left(x_{s}\right)}
$$

We will use the following obvious fact:
Fact 1. If $x$ does not contain digits smaller than $k$, then $f_{i}(x)=4^{f_{i+1}(x)}$ for all $i=0, \ldots, k-1$. In particular, $f_{i}(\varepsilon)=4^{9-i}$ for all $i=0,1, \ldots, 9$.
Moreover, by induction on $k=9,8, \ldots, 0$ it follows easily:
Fact 2. If the nonempty string $x$ does not contain digits smaller than $k$, then $f_{i}(x)>f_{i}(\varepsilon)$ for all $i=0, \ldots, k$.
We will show the essential fact:
Fact 3. $f_{0}(n)>f_{0}(h(n))$.
Then the empty string will necessarily be reached after a finite number of applications of $h$. But starting from a string without leading zeros, $\varepsilon$ can only be reached via the strings $1 \rightarrow 00 \rightarrow 0 \rightarrow \varepsilon$. Hence also the number 1 will appear after a finite number of applications of $h$.
Proof of Fact 3. If the last digit $r$ of $n$ is 0 , then we write $n=x_{0} 0 \ldots 0 x_{m-1} 0 \varepsilon$ where the $x_{i}$ do not contain the digit 0 . Then $h(n)=x_{0} 0 \ldots 0 x_{m-1}$ and $f_{0}(n)-f_{0}(h(n))=f_{0}(\varepsilon)>0$.
So let the last digit $r$ of $n$ be at least 1 . Let $L=y k$ and $R=z r$ be the corresponding left and right parts where $y$ is some string, $k \leq r-1$ and the string $z$ consists only of digits not less
than $r$. Then $n=y k z r$ and $h(n)=y k z(r-1) z(r-1)$. Let $d(y)$ be the smallest digit of $y$. We consider two cases which do not exclude each other.

Case 1. $d(y) \geq k$.
Then

$$
f_{k}(n)-f_{k}(h(n))=f_{k}(z r)-f_{k}(z(r-1) z(r-1)) .
$$

In view of Fact 1 this difference is positive if and only if

$$
f_{r-1}(z r)-f_{r-1}(z(r-1) z(r-1))>0 .
$$

We have, using Fact 2,

$$
f_{r-1}(z r)=4^{f_{r}(z r)}=4^{f_{r}(z)+4^{f_{r+1}(z)}} \geq 4 \cdot 4^{f_{r}(z)}>4^{f_{r}(z)}+4^{f_{r}(z)}+4^{f_{r}(\varepsilon)}=f_{r-1}(z(r-1) z(r-1)) .
$$

Here we use the additional definition $f_{10}(\varepsilon)=0$ if $r=9$. Consequently, $f_{k}(n)-f_{k}(h(n))>0$ and according to Fact $1, f_{0}(n)-f_{0}(h(n))>0$.
Case 2. $d(y) \leq k$.
We prove by induction on $d(y)=k, k-1, \ldots, 0$ that $f_{i}(n)-f_{i}(h(n))>0$ for all $i=0, \ldots, d(y)$. By Fact 1, it suffices to do so for $i=d(y)$. The initialization $d(y)=k$ was already treated in Case 1. Let $t=d(y)<k$. Write $y$ in the form utv where $v$ does not contain digits $\leq t$. Then, in view of the induction hypothesis,

$$
f_{t}(n)-f_{t}(h(n))=f_{t}(v k z r)-f_{t}(v k z(r-1) z(r-1))=4^{f_{t+1}(v k z r)}-4^{f_{t+1}(v k z(r-1) z(r-1))}>0 .
$$

Thus the inequality $f_{d(y)}(n)-f_{d(y)}(h(n))>0$ is established and from Fact 1 it follows that $f_{0}(n)-f_{0}(h(n))>0$.

Solution 2. We identify integers $n \geq 2$ with the digit-strings, briefly strings, of their decimal representation and extend the definition of $h$ to all non-empty strings with digits from 0 to 9. Moreover, let us define that the empty string, $\varepsilon$, is being mapped to the empty string. In the following all functions map the set of strings into the set of strings. For two functions $f$ and $g$ let $g \circ f$ be defined by $(g \circ f)(x)=g(f(x))$ for all strings $x$ and let, for non-negative integers $n, f^{n}$ denote the $n$-fold application of $f$. For any string $x$ let $s(x)$ be the smallest digit of $x$, and for the empty string let $s(\varepsilon)=\infty$. We define nine functions $g_{1}, \ldots, g_{9}$ as follows: Let $k \in\{1, \ldots, 9\}$ and let $x$ be a string. If $x=\varepsilon$ then $g_{k}(x)=\varepsilon$. Otherwise, write $x$ in the form $x=y z r$ where $y$ is either the empty string or ends with a digit smaller than $k, s(z) \geq k$ and $r$ is the rightmost digit of $x$. Then $g_{k}(x)=z r$.
Lemma 1. We have $g_{k} \circ h=g_{k} \circ h \circ g_{k}$ for all $k=1, \ldots, 9$.
Proof of Lemma 1. Let $x=y z r$ be as in the definition of $g_{k}$. If $y=\varepsilon$, then $g_{k}(x)=x$, whence

$$
\begin{equation*}
g_{k}(h(x))=g_{k}\left(h\left(g_{k}(x)\right) .\right. \tag{1}
\end{equation*}
$$

So let $y \neq \varepsilon$.
Case 1. $z$ contains a digit smaller than $r$.
Let $z=u a v$ where $a<r$ and $s(v) \geq r$. Then

$$
h(x)= \begin{cases}\text { yuav } & \text { if } r=0, \\ \operatorname{yuav}(r-1) v(r-1) & \text { if } r>0\end{cases}
$$

and

$$
h\left(g_{k}(x)\right)=h(z r)=h(u a v r)= \begin{cases}\text { uav } & \text { if } r=0 \\ \operatorname{uav}(r-1) v(r-1) & \text { if } r>0\end{cases}
$$

Since $y$ ends with a digit smaller than $k,(1)$ is obviously true.
Case 2. $z$ does not contain a digit smaller than $r$.
Let $y=u v$ where $u$ is either the empty string or ends with a digit smaller than $r$ and $s(v) \geq r$. We have

$$
h(x)= \begin{cases}u v z & \text { if } r=0 \\ u v z(r-1) v z(r-1) & \text { if } r>0\end{cases}
$$

and

$$
h\left(g_{k}(x)\right)=h(z r)= \begin{cases}z & \text { if } r=0 \\ z(r-1) z(r-1) & \text { if } r>0\end{cases}
$$

Recall that $y$ and hence $v$ ends with a digit smaller than $k$, but all digits of $v$ are at least $r$. Now if $r>k$, then $v=\varepsilon$, whence the terminal digit of $u$ is smaller than $k$, which entails

$$
g_{k}(h(x))=z(r-1) z(r-1)=g_{k}\left(h\left(g_{k}(x)\right)\right) .
$$

If $r \leq k$, then

$$
g_{k}(h(x))=z(r-1)=g_{k}\left(h\left(g_{k}(x)\right)\right),
$$

so that in both cases (1) is true. Thus Lemma 1 is proved.
Lemma 2. Let $k \in\{1, \ldots, 9\}$, let $x$ be a non-empty string and let $n$ be a positive integer. If $h^{n}(x)=\varepsilon$ then $\left(g_{k} \circ h\right)^{n}(x)=\varepsilon$.
Proof of Lemma 2. We proceed by induction on $n$. If $n=1$ we have

$$
\varepsilon=h(x)=g_{k}(h(x))=\left(g_{k} \circ h\right)(x) .
$$

Now consider the step from $n-1$ to $n$ where $n \geq 2$. Let $h^{n}(x)=\varepsilon$ and let $y=h(x)$. Then $h^{n-1}(y)=\varepsilon$ and by the induction hypothesis $\left(g_{k} \circ h\right)^{n-1}(y)=\varepsilon$. In view of Lemma 1 ,

$$
\begin{aligned}
& \varepsilon=\left(g_{k} \circ h\right)^{n-2}\left(\left(g_{k} \circ h\right)(y)\right)=\left(g_{k} \circ h\right)^{n-2}\left(g_{k}(h(y))\right. \\
&=\left(g_{k} \circ h\right)^{n-2}\left(g_{k}\left(h\left(g_{k}(y)\right)\right)=\left(g_{k} \circ h\right)^{n-2}\left(g_{k}\left(h\left(g_{k}(h(x))\right)\right)=\left(g_{k} \circ h\right)^{n}(x)\right.\right.
\end{aligned}
$$

Thus the induction step is complete and Lemma 2 is proved.
We say that the non-empty string $x$ terminates if $h^{n}(x)=\varepsilon$ for some non-negative integer $n$.
Lemma 3. Let $x=y z r$ where $s(y) \geq k, s(z) \geq k, y$ ends with the digit $k$ and $z$ is possibly empty. If $y$ and $z r$ terminate then also $x$ terminates.
Proof of Lemma 3. Suppose that $y$ and $z r$ terminate. We proceed by induction on $k$. Let $k=0$. Obviously, $h(y w)=y h(w)$ for any non-empty string $w$. Let $h^{n}(z r)=\epsilon$. It follows easily by induction on $m$ that $h^{m}(y z r)=y h^{m}(z r)$ for $m=1, \ldots, n$. Consequently, $h^{n}(y z r)=y$. Since $y$ terminates, also $x=y z r$ terminates.
Now let the assertion be true for all nonnegative integers less than $k$ and let us prove it for $k$ where $k \geq 1$. It turns out that it is sufficient to prove that $y g_{k}(h(z r))$ terminates. Indeed:
Case 1. $r=0$.
Then $h(y z r)=y z=y g_{k}(h(z r))$.
Case 2. $0<r \leq k$.
We have $h(z r)=z(r-1) z(r-1)$ and $g_{k}(h(z r))=z(r-1)$. Then $h(y z r)=y z(r-1) y z(r-$
$1)=y g_{k}(h(z r)) y g_{k}(h(z r))$ and we may apply the induction hypothesis to see that if $\left.y g_{k} h(z r)\right)$ terminates, then $h(y z r)$ terminates.

Case 3. $r>k$.
Then $h(y z r)=y h(z r)=y g_{k}(h(z r))$.
Note that $y g_{k}(h(z r))$ has the form $y z^{\prime} r^{\prime}$ where $s\left(z^{\prime}\right) \geq k$. By the same arguments it is sufficient to prove that $y g_{k}\left(h\left(z^{\prime} r^{\prime}\right)\right)=y\left(g_{k} \circ h\right)^{2}(z r)$ terminates and, by induction, that $y\left(g_{k} \circ h\right)^{m}(z r)$ terminates for some positive integer $m$. In view of Lemma 2 there is some $m$ such that ( $g_{k} \circ$ $h)^{m}(z r)=\epsilon$, so $x=y z r$ terminates if $y$ terminates. Thus Lemma 3 is proved.
Now assume that there is some string $x$ that does not terminate. We choose $x$ minimal. If $x \geq 10$, we can write $x$ in the form $x=y z r$ of Lemma 3 and by this lemma $x$ terminates since $y$ and $z r$ are smaller than $x$. If $x \leq 9$, then $h(x)=(x-1)(x-1)$ and $h(x)$ terminates again by Lemma 3 and the minimal choice of $x$.

Solution 3. We commence by introducing some terminology. Instead of integers, we will consider the set $S$ of all strings consisting of the digits $0,1, \ldots, 9$, including the empty string $\epsilon$. If $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ is a nonempty string, we let $\rho(a)=a_{n}$ denote the terminal digit of $a$ and $\lambda(a)$ be the string with the last digit removed. We also define $\lambda(\epsilon)=\epsilon$ and denote the set of non-negative integers by $\mathbb{N}_{0}$.
Now let $k \in\{0,1,2, \ldots, 9\}$ denote any digit. We define a function $f_{k}: S \longrightarrow S$ on the set of strings: First, if the terminal digit of $n$ belongs to $\{0,1, \ldots, k\}$, then $f_{k}(n)$ is obtained from $n$ by deleting this terminal digit, i.e $f_{k}(n)=\lambda(n)$. Secondly, if the terminal digit of $n$ belongs to $\{k+1, \ldots, 9\}$, then $f_{k}(n)$ is obtained from $n$ by the process described in the problem. We also define $f_{k}(\epsilon)=\epsilon$. Note that up to the definition for integers $n \leq 1$, the function $f_{0}$ coincides with the function $h$ in the problem, through interpreting integers as digit strings. The argument will be roughly as follows. We begin by introducing a straightforward generalization of our claim about $f_{0}$. Then it will be easy to see that $f_{9}$ has all these stronger properties, which means that is suffices to show for $k \in\{0,1, \ldots, 8\}$ that $f_{k}$ possesses these properties provided that $f_{k+1}$ does.
We continue to use $k$ to denote any digit. The operation $f_{k}$ is said to be separating, if the followings holds: Whenever $a$ is an initial segment of $b$, there is some $N \in \mathbb{N}_{0}$ such that $f_{k}^{N}(b)=a$. The following two notions only apply to the case where $f_{k}$ is indeed separating, otherwise they remain undefined. For every $a \in S$ we denote the least $N \in \mathbb{N}_{0}$ for which $f_{k}^{N}(a)=\epsilon$ occurs by $g_{k}(a)$ (because $\epsilon$ is an initial segment of $a$, such an $N$ exists if $f_{k}$ is separating). If for every two strings $a$ and $b$ such that $a$ is a terminal segment of $b$ one has $g_{k}(a) \leq g_{k}(b)$, we say that $f_{k}$ is coherent. In case that $f_{k}$ is separating and coherent we call the digit $k$ seductive.
As $f_{9}(a)=\lambda(a)$ for all $a$, it is obvious that 9 is seductive. Hence in order to show that 0 is seductive, which clearly implies the statement of the problem, it suffices to take any $k \in\{0,1, \ldots, 8\}$ such that $k+1$ is seductive and to prove that $k$ has to be seductive as well. Note that in doing so, we have the function $g_{k+1}$ at our disposal. We have to establish two things and we begin with

Step 1. $f_{k}$ is separating.

Before embarking on the proof of this, we record a useful observation which is easily proved by induction on $M$.

Claim 1. For any strings $A, B$ and any positive integer $M$ such that $f_{k}^{M-1}(B) \neq \epsilon$, we have

$$
f_{k}^{M}(A k B)=A k f_{k}^{M}(B)
$$

Now we call a pair $(a, b)$ of strings wicked provided that $a$ is an initial segment of $b$, but there is no $N \in \mathbb{N}_{0}$ such that $f_{k}^{N}(b)=a$. We need to show that there are none, so assume that there were such pairs. Choose a wicked pair $(a, b)$ for which $g_{k+1}(b)$ attains its minimal possible value. Obviously $b \neq \epsilon$ for any wicked pair $(a, b)$. Let $z$ denote the terminal digit of $b$. Observe that $a \neq b$, which means that $a$ is also an initial segment of $\lambda(b)$. To facilitate the construction of the eventual contradiction, we prove
Claim 2. There cannot be an $N \in \mathbb{N}_{0}$ such that

$$
f_{k}^{N}(b)=\lambda(b)
$$

Proof of Claim 2. For suppose that such an $N$ existed. Because $g_{k+1}(\lambda(b))<g_{k+1}(b)$ in view of the coherency of $f_{k+1}$, the pair $(a, \lambda(b))$ is not wicked. But then there is some $N^{\prime}$ for which $f_{k}^{N^{\prime}}(\lambda(b))=a$ which entails $f_{k}^{N+N^{\prime}}(b)=a$, contradiction. Hence Claim 2 is proved.

It follows that $z \leq k$ is impossible, for otherwise $N=1$ violated Claim 2.
Also $z>k+1$ is impossible: Set $B=f_{k}(b)$. Then also $f_{k+1}(b)=B$, but $g_{k+1}(B)<g_{k+1}(b)$ and $a$ is an initial segment of $B$. Thus the pair $(a, B)$ is not wicked. Hence there is some $N \in \mathbb{N}_{0}$ with $a=f_{k}^{N}(B)$, which, however, entails $a=f_{k}^{N+1}(b)$.
We are left with the case $z=k+1$. Let $L$ denote the left part and $R=R^{*}(k+1)$ the right part of $b$. Then we have symbolically

$$
f_{k}(b)=L R^{*} k R^{*} k, f_{k}^{2}(b)=L R^{*} k R^{*} \quad \text { and } \quad f_{k+1}(b)=L R^{*} .
$$

Using that $R^{*}$ is a terminal segment of $L R^{*}$ and the coherency of $f_{k+1}$, we infer

$$
g_{k+1}\left(R^{*}\right) \leq g_{k+1}\left(L R^{*}\right)<g_{k+1}(b) .
$$

Hence the pair ( $\epsilon, R^{*}$ ) is not wicked, so there is some minimal $M \in \mathbb{N}_{0}$ with $f_{k}^{M}\left(R^{*}\right)=\epsilon$ and by Claim 1 it follows that $f_{k}^{2+M}(b)=L R^{*} k$. Finally, we infer that $\lambda(b)=L R^{*}=f_{k}\left(L R^{*} k\right)=$ $f_{k}^{3+M}(b)$, which yields a contradiction to Claim 2.
This final contradiction establishes that $f_{k}$ is indeed separating.

Step 2. $f_{k}$ is coherent.

To prepare the proof of this, we introduce some further pieces of terminology. A nonempty string $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ is called a hypostasis, if $a_{n}<a_{i}$ for all $i=1, \ldots, n-1$. Reading an arbitrary string $a$ backwards, we easily find a, possibly empty, sequence $\left(A_{1}, A_{2}, \ldots, A_{m}\right)$ of hypostases such that $\rho\left(A_{1}\right) \leq \rho\left(A_{2}\right) \leq \cdots \leq \rho\left(A_{m}\right)$ and, symbolically, $a=A_{1} A_{2} \ldots A_{m}$. The latter sequence is referred to as the decomposition of $a$. So, for instance, $(20,0,9)$ is the decomposition of 2009 and the string 50 is a hypostasis. Next we explain when we say about two strings $a$ and $b$ that $a$ is injectible into $b$. The definition is by induction on the length of $b$. Let $\left(B_{1}, B_{2}, \ldots, B_{n}\right)$ be the decomposition of $b$ into hypostases. Then $a$ is injectible into $b$ if for the decomposition $\left(A_{1}, A_{2}, \ldots, A_{m}\right)$ of $a$ there is a strictly increasing function $H:\{1,2, \ldots, m\} \longrightarrow\{1,2, \ldots, n\}$ satisfying

$$
\rho\left(A_{i}\right)=\rho\left(B_{H(i)}\right) \text { for all } i=1, \ldots, m \text {; }
$$

$\lambda\left(A_{i}\right)$ is injectible into $\lambda\left(B_{H(i)}\right)$ for all $i=1, \ldots, m$.
If one can choose $H$ with $H(m)=n$, then we say that $a$ is strongly injectible into $b$. Obviously, if $a$ is a terminal segment of $b$, then $a$ is strongly injectible into $b$.

Claim 3. If $a$ and $b$ are two nonempty strings such that $a$ is strongly injectible into $b$, then $\lambda(a)$ is injectible into $\lambda(b)$.

Proof of Claim 3. Let $\left(B_{1}, B_{2}, \ldots, B_{n}\right)$ be the decomposition of $b$ and let $\left(A_{1}, A_{2}, \ldots, A_{m}\right)$ be the decomposition of $a$. Take a function $H$ exemplifying that $a$ is strongly injectible into $b$. Let $\left(C_{1}, C_{2}, \ldots, C_{r}\right)$ be the decomposition of $\lambda\left(A_{m}\right)$ and let $\left(D_{1}, D_{2}, \ldots, D_{s}\right)$ be the decomposition of $\lambda\left(B_{n}\right)$. Choose a strictly increasing $H^{\prime}:\{1,2, \ldots, r\} \longrightarrow\{1,2, \ldots s\}$ witnessing that $\lambda\left(A_{m}\right)$ is injectible into $\lambda\left(B_{n}\right)$. Clearly, $\left(A_{1}, A_{2}, \ldots, A_{m-1}, C_{1}, C_{2}, \ldots, C_{r}\right)$ is the decomposition of $\lambda(a)$ and $\left(B_{1}, B_{2}, \ldots, B_{n-1}, D_{1}, D_{2}, \ldots, D_{s}\right)$ is the decomposition of $\lambda(b)$. Then the function $H^{\prime \prime}:\{1,2, \ldots, m+r-1\} \longrightarrow\{1,2, \ldots, n+s-1\}$ given by $H^{\prime \prime}(i)=H(i)$ for $i=1,2, \ldots, m-1$ and $H^{\prime \prime}(m-1+i)=n-1+H^{\prime}(i)$ for $i=1,2, \ldots, r$ exemplifies that $\lambda(a)$ is injectible into $\lambda(b)$, which finishes the proof of the claim.

A pair $(a, b)$ of strings is called aggressive if $a$ is injectible into $b$ and nevertheless $g_{k}(a)>g_{k}(b)$. Observe that if $f_{k}$ was incoherent, which we shall assume from now on, then such pairs existed. Now among all aggressive pairs we choose one, say $(a, b)$, for which $g_{k}(b)$ attains its least possible value. Obviously $f_{k}(a)$ cannot be injectible into $f_{k}(b)$, for otherwise the pair $\left(f_{k}(a), f_{k}(b)\right)$ was aggressive and contradicted our choice of $(a, b)$. Let $\left(A_{1}, A_{2}, \ldots, A_{m}\right)$ and $\left(B_{1}, B_{2}, \ldots, B_{n}\right)$ be the decompositions of $a$ and $b$ and take a function $H:\{1,2, \ldots, m\} \longrightarrow\{1,2, \ldots, n\}$ exemplifying that $a$ is indeed injectible into $b$. If we had $H(m)<n$, then $a$ was also injectible into the number $b^{\prime}$ whose decomposition is $\left(B_{1}, B_{2}, \ldots, B_{n-1}\right)$ and by separativity of $f_{k}$ we obtained $g_{k}\left(b^{\prime}\right)<g_{k}(b)$, whence the pair ( $a, b^{\prime}$ ) was also aggressive, contrary to the minimality condition imposed on $b$. Therefore $a$ is strongly injectible into $b$. In particular, $a$ and $b$ have a common terminal digit, say $z$. If we had $z \leq k$, then $f_{k}(a)=\lambda(a)$ and $f_{k}(b)=\lambda(b)$, so that by Claim $3, f_{k}(a)$ was injectible into $f_{k}(b)$, which is a contradiction. Hence, $z \geq k+1$.
Now let $r$ be the minimal element of $\{1,2, \ldots, m\}$ for which $\rho\left(A_{r}\right)=z$. Then the maximal right part of $a$ consisting of digits $\geq z$ is equal to $R_{a}$, the string whose decomposition is $\left(A_{r}, A_{r+1}, \ldots, A_{m}\right)$. Then $R_{a}-1$ is a hypostasis and $\left(A_{1}, \ldots, A_{r-1}, R_{a}-1, R_{a}-1\right)$ is the decomposition of $f_{k}(a)$. Defining $s$ and $R_{b}$ in a similar fashion with respect to $b$, we see that $\left(B_{1}, \ldots, B_{s-1}, R_{b}-1, R_{b}-1\right)$ is the decomposition of $f_{k}(b)$. The definition of injectibility then easily entails that $R_{a}$ is strongly injectible into $R_{b}$. It follows from Claim 3 that $\lambda\left(R_{a}\right)=$ $\lambda\left(R_{a}-1\right)$ is injectible into $\lambda\left(R_{b}\right)=\lambda\left(R_{b}-1\right)$, whence the function $H^{\prime}:\{1,2, \ldots, r+1\} \longrightarrow$ $\{1,2, \ldots, s+1\}$, given by $H^{\prime}(i)=H(i)$ for $i=1,2, \ldots, r-1, H^{\prime}(r)=s$ and $H^{\prime}(r+1)=s+1$ exemplifies that $f_{k}(a)$ is injectible into $f_{k}(b)$, which yields a contradiction as before.
This shows that aggressive pairs cannot exist, whence $f_{k}$ is indeed coherent, which finishes the proof of the seductivity of $k$, whereby the problem is finally solved.

## Geometry

## G1 BEL (Belgium)

Let $A B C$ be a triangle with $A B=A C$. The angle bisectors of $A$ and $B$ meet the sides $B C$ and $A C$ in $D$ and $E$, respectively. Let $K$ be the incenter of triangle $A D C$. Suppose that $\angle B E K=45^{\circ}$. Find all possible values of $\angle B A C$.

Solution 1. Answer: $\angle B A C=60^{\circ}$ or $\angle B A C=90^{\circ}$ are possible values and the only possible values.

Let $I$ be the incenter of triangle $A B C$, then $K$ lies on the line $C I$. Let $F$ be the point, where the incircle of triangle $A B C$ touches the side $A C$; then the segments $I F$ and $I D$ have the same length and are perpendicular to $A C$ and $B C$, respectively.


Figure 1


Figure 2

Let $P, Q$ and $R$ be the points where the incircle of triangle $A D C$ touches the sides $A D, D C$ and $C A$, respectively. Since $K$ and $I$ lie on the angle bisector of $\angle A C D$, the segments $I D$ and $I F$ are symmetric with respect to the line $I C$. Hence there is a point $S$ on $I F$ where the incircle of triangle $A D C$ touches the segment $I F$. Then segments $K P, K Q, K R$ and $K S$ all have the same length and are perpendicular to $A D, D C, C A$ and $I F$, respectively. So - regardless of the value of $\angle B E K$ - the quadrilateral $K R F S$ is a square and $\angle S F K=\angle K F C=45^{\circ}$.
Consider the case $\angle B A C=60^{\circ}$ (see Figure 1). Then triangle $A B C$ is equilateral. Furthermore we have $F=E$, hence $\angle B E K=\angle I F K=\angle S E K=45^{\circ}$. So $60^{\circ}$ is a possible value for $\angle B A C$.
Now consider the case $\angle B A C=90^{\circ}$ (see Figure 2). Then $\angle C B A=\angle A C B=45^{\circ}$. Furthermore, $\angle K I E=\frac{1}{2} \angle C B A+\frac{1}{2} \angle A C B=45^{\circ}, \angle A E B=180^{\circ}-90^{\circ}-22.5^{\circ}=67.5^{\circ}$ and $\angle E I A=\angle B I D=180^{\circ}-90^{\circ}-22.5^{\circ}=67.5^{\circ}$. Hence triangle $I E A$ is isosceles and a reflection of the bisector of $\angle I A E$ takes $I$ to $E$ and $K$ to itself. So triangle $I K E$ is symmetric with respect to this axis, i.e. $\angle K I E=\angle I E K=\angle B E K=45^{\circ}$. So $90^{\circ}$ is a possible value for $\angle B A C$, too.
If, on the other hand, $\angle B E K=45^{\circ}$ then $\angle B E K=\angle I E K=\angle I F K=45^{\circ}$. Then

- either $F=E$, which makes the angle bisector $B I$ be an altitude, i.e., which makes triangle $A B C$ isosceles with base $A C$ and hence equilateral and so $\angle B A C=60^{\circ}$,
- or $E$ lies between $F$ and $C$, which makes the points $K, E, F$ and $I$ concyclic, so $45^{\circ}=$ $\angle K F C=\angle K F E=\angle K I E=\angle C B I+\angle I C B=2 \cdot \angle I C B=90^{\circ}-\frac{1}{2} \angle B A C$, and so $\angle B A C=90^{\circ}$,
- or $F$ lies between $E$ and $C$, then again, $K, E, F$ and $I$ are concyclic, so $45^{\circ}=\angle K F C=$ $180^{\circ}-\angle K F E=\angle K I E$, which yields the same result $\angle B A C=90^{\circ}$. (However, for $\angle B A C=90^{\circ} E$ lies, in fact, between $F$ and $C$, see Figure 2. So this case does not occur.)
This proves $90^{\circ}$ and $60^{\circ}$ to be the only possible values for $\angle B A C$.

Solution 2. Denote angles at $A, B$ and $C$ as usual by $\alpha, \beta$ and $\gamma$. Since triangle $A B C$ is isosceles, we have $\beta=\gamma=90^{\circ}-\frac{\alpha}{2}<90^{\circ}$, so $\angle E C K=45^{\circ}-\frac{\alpha}{4}=\angle K C D$. Since $K$ is the incenter of triangle $A D C$, we have $\angle C D K=\angle K D A=45^{\circ}$; furthermore $\angle D I C=45^{\circ}+\frac{\alpha}{4}$. Now, if $\angle B E K=45^{\circ}$, easy calculations within triangles $B C E$ and $K C E$ yield

$$
\begin{aligned}
& \angle K E C=180^{\circ}-\frac{\beta}{2}-45^{\circ}-\beta=135^{\circ}-\frac{3}{2} \beta=\frac{3}{2}\left(90^{\circ}-\beta\right)=\frac{3}{4} \alpha, \\
& \angle I K E=\frac{3}{4} \alpha+45^{\circ}-\frac{\alpha}{4}=45^{\circ}+\frac{\alpha}{2} .
\end{aligned}
$$

So in triangles $I C E, I K E, I D K$ and $I D C$ we have (see Figure 3)

$$
\begin{array}{ll}
\frac{I C}{I E}=\frac{\sin \angle I E C}{\sin \angle E C I}=\frac{\sin \left(45^{\circ}+\frac{3}{4} \alpha\right)}{\sin \left(45^{\circ}-\frac{\alpha}{4}\right)}, & \frac{I E}{I K}=\frac{\sin \angle E K I}{\sin \angle I E K}=\frac{\sin \left(45^{\circ}+\frac{\alpha}{2}\right)}{\sin 45^{\circ}} \\
\frac{I K}{I D}=\frac{\sin \angle K D I}{\sin \angle I K D}=\frac{\sin 45^{\circ}}{\sin \left(90^{\circ}-\frac{\alpha}{4}\right)}, & \frac{I D}{I C}=\frac{\sin \angle I C D}{\sin \angle C D I}=\frac{\sin \left(45^{\circ}-\frac{\alpha}{4}\right)}{\sin 90^{\circ}}
\end{array}
$$



Figure 3
Multiplication of these four equations yields

$$
1=\frac{\sin \left(45^{\circ}+\frac{3}{4} \alpha\right) \sin \left(45^{\circ}+\frac{\alpha}{2}\right)}{\sin \left(90^{\circ}-\frac{\alpha}{4}\right)}
$$

But, since

$$
\begin{aligned}
\sin \left(90^{\circ}-\frac{\alpha}{4}\right) & =\cos \frac{\alpha}{4}=\cos \left(\left(45^{\circ}+\frac{3}{4} \alpha\right)-\left(45^{\circ}+\frac{\alpha}{2}\right)\right) \\
& =\cos \left(45^{\circ}+\frac{3}{4} \alpha\right) \cos \left(45^{\circ}+\frac{\alpha}{2}\right)+\sin \left(45^{\circ}+\frac{3}{4} \alpha\right) \sin \left(45^{\circ}+\frac{\alpha}{2}\right),
\end{aligned}
$$

this is equivalent to

$$
\sin \left(45^{\circ}+\frac{3}{4} \alpha\right) \sin \left(45^{\circ}+\frac{\alpha}{2}\right)=\cos \left(45^{\circ}+\frac{3}{4} \alpha\right) \cos \left(45^{\circ}+\frac{\alpha}{2}\right)+\sin \left(45^{\circ}+\frac{3}{4} \alpha\right) \sin \left(45^{\circ}+\frac{\alpha}{2}\right)
$$

and finally

$$
\cos \left(45^{\circ}+\frac{3}{4} \alpha\right) \cos \left(45^{\circ}+\frac{\alpha}{2}\right)=0
$$

But this means $\cos \left(45^{\circ}+\frac{3}{4} \alpha\right)=0$, hence $45^{\circ}+\frac{3}{4} \alpha=90^{\circ}$, i.e. $\alpha=60^{\circ}$ or $\cos \left(45^{\circ}+\frac{\alpha}{2}\right)=0$, hence $45^{\circ}+\frac{\alpha}{2}=90^{\circ}$, i.e. $\alpha=90^{\circ}$. So these values are the only two possible values for $\alpha$.
On the other hand, both $\alpha=90^{\circ}$ and $\alpha=60^{\circ}$ yield $\angle B E K=45^{\circ}$, this was shown in Solution 1.

## G2 RUS (Russian Federation)

Let $A B C$ be a triangle with circumcenter $O$. The points $P$ and $Q$ are interior points of the sides $C A$ and $A B$, respectively. The circle $k$ passes through the midpoints of the segments $B P$, $C Q$, and $P Q$. Prove that if the line $P Q$ is tangent to circle $k$ then $O P=O Q$.

Solution 1. Let $K, L, M, B^{\prime}, C^{\prime}$ be the midpoints of $B P, C Q, P Q, C A$, and $A B$, respectively (see Figure 1). Since $C A \| L M$, we have $\angle L M P=\angle Q P A$. Since $k$ touches the segment $P Q$ at $M$, we find $\angle L M P=\angle L K M$. Thus $\angle Q P A=\angle L K M$. Similarly it follows from $A B \| M K$ that $\angle P Q A=\angle K L M$. Therefore, triangles $A P Q$ and $M K L$ are similar, hence

$$
\begin{equation*}
\frac{A P}{A Q}=\frac{M K}{M L}=\frac{\frac{Q B}{2}}{\frac{P C}{2}}=\frac{Q B}{P C} \tag{1}
\end{equation*}
$$

Now (1) is equivalent to $A P \cdot P C=A Q \cdot Q B$ which means that the power of points $P$ and $Q$ with respect to the circumcircle of $\triangle A B C$ are equal, hence $O P=O Q$.


Figure 1

Comment. The last argument can also be established by the following calculation:

$$
\begin{aligned}
O P^{2}-O Q^{2} & =O B^{\prime 2}+B^{\prime} P^{2}-O C^{\prime 2}-C^{\prime} Q^{2} \\
& =\left(O A^{2}-A B^{\prime 2}\right)+B^{\prime} P^{2}-\left(O A^{2}-A C^{\prime 2}\right)-C^{\prime} Q^{2} \\
& =\left(A C^{\prime 2}-C^{\prime} Q^{2}\right)-\left(A B^{\prime 2}-B^{\prime} P^{2}\right) \\
& =\left(A C^{\prime}-C^{\prime} Q\right)\left(A C^{\prime}+C^{\prime} Q\right)-\left(A B^{\prime}-B^{\prime} P\right)\left(A B^{\prime}+B^{\prime} P\right) \\
& =A Q \cdot Q B-A P \cdot P C .
\end{aligned}
$$

With (1), we conclude $O P^{2}-O Q^{2}=0$, as desired.

Solution 2. Again, denote by $K, L, M$ the midpoints of segments $B P, C Q$, and $P Q$, respectively. Let $O, S, T$ be the circumcenters of triangles $A B C, K L M$, and $A P Q$, respectively (see Figure 2). Note that $M K$ and $L M$ are the midlines in triangles $B P Q$ and $C P Q$, respectively, so $\overrightarrow{M K}=\frac{1}{2} \overrightarrow{Q B}$ and $\overrightarrow{M L}=\frac{1}{2} \overrightarrow{P C}$. Denote by $\operatorname{pr}_{l}(\vec{v})$ the projection of vector $\vec{v}$ onto line $l$. Then $\operatorname{pr}_{A B}(\overrightarrow{O T})=\operatorname{pr}_{A B}(\overrightarrow{O A}-\overrightarrow{T A})=\frac{1}{2} \overrightarrow{B A}-\frac{1}{2} \overrightarrow{Q A}=\frac{1}{2} \overrightarrow{B Q}=\overrightarrow{K M}$ and $\operatorname{pr}_{A B}(\overrightarrow{S M})=\operatorname{pr}_{M K}(\overrightarrow{S M})=$ $\frac{1}{2} \overrightarrow{K M}=\frac{1}{2} \operatorname{pr}_{A B}(\overrightarrow{O T})$. Analogously we get $\operatorname{pr}_{C A}(\overrightarrow{S M})=\frac{1}{2} \operatorname{pr}_{C A}(\overrightarrow{O T})$. Since $A B$ and $C A$ are not parallel, this implies that $\overrightarrow{S M}=\frac{1}{2} \overrightarrow{O T}$.


Figure 2
Now, since the circle $k$ touches $P Q$ at $M$, we get $S M \perp P Q$, hence $O T \perp P Q$. Since $T$ is equidistant from $P$ and $Q$, the line $O T$ is a perpendicular bisector of segment $P Q$, and hence $O$ is equidistant from $P$ and $Q$ which finishes the proof.

## G3 IRN (Islamic Republic of Iran)

Let $A B C$ be a triangle. The incircle of $A B C$ touches the sides $A B$ and $A C$ at the points $Z$ and $Y$, respectively. Let $G$ be the point where the lines $B Y$ and $C Z$ meet, and let $R$ and $S$ be points such that the two quadrilaterals $B C Y R$ and $B C S Z$ are parallelograms.
Prove that $G R=G S$.

Solution 1. Denote by $k$ the incircle and by $k_{a}$ the excircle opposite to $A$ of triangle $A B C$. Let $k$ and $k_{a}$ touch the side $B C$ at the points $X$ and $T$, respectively, let $k_{a}$ touch the lines $A B$ and $A C$ at the points $P$ and $Q$, respectively. We use several times the fact that opposing sides of a parallelogram are of equal length, that points of contact of the excircle and incircle to a side of a triangle lie symmetric with respect to the midpoint of this side and that segments on two tangents to a circle defined by the points of contact and their point of intersection have the same length. So we conclude

$$
\begin{gathered}
Z P=Z B+B P=X B+B T=B X+C X=Z S \text { and } \\
C Q=C T=B X=B Z=C S .
\end{gathered}
$$



So for each of the points $Z, C$, their distances to $S$ equal the length of a tangent segment from this point to $k_{a}$. It is well-known, that all points with this property lie on the line $Z C$, which is the radical axis of $S$ and $k_{a}$. Similar arguments yield that $B Y$ is the radical axis of $R$ and $k_{a}$. So the point of intersection of $Z C$ and $B Y$, which is $G$ by definition, is the radical center of $R, S$ and $k_{a}$, from which the claim $G R=G S$ follows immediately.

Solution 2. Denote $x=A Z=A Y, y=B Z=B X, z=C X=C Y, p=Z G, q=G C$. Several lengthy calculations (Menelaos' theorem in triangle $A Z C$, law of Cosines in triangles $A B C$ and $A Z C$ and Stewart's theorem in triangle $Z C S$ ) give four equations for $p, q, \cos \alpha$
and $G S$ in terms of $x, y$, and $z$ that can be resolved for $G S$. The result is symmetric in $y$ and $z$, so $G R=G S$. More in detail this means:
The line $B Y$ intersects the sides of triangle $A Z C$, so Menelaos' theorem yields $\frac{p}{q} \cdot \frac{z}{x} \cdot \frac{x+y}{y}=1$, hence

$$
\begin{equation*}
\frac{p}{q}=\frac{x y}{y z+z x} . \tag{1}
\end{equation*}
$$

Since we only want to show that the term for $G S$ is symmetric in $y$ and $z$, we abbreviate terms that are symmetric in $y$ and $z$ by capital letters, starting with $N=x y+y z+z x$. So (1) implies

$$
\begin{equation*}
\frac{p}{p+q}=\frac{x y}{x y+y z+z x}=\frac{x y}{N} \quad \text { and } \quad \frac{q}{p+q}=\frac{y z+z x}{x y+y z+z x}=\frac{y z+z x}{N} . \tag{2}
\end{equation*}
$$

Now the law of Cosines in triangle $A B C$ yields

$$
\cos \alpha=\frac{(x+y)^{2}+(x+z)^{2}-(y+z)^{2}}{2(x+y)(x+z)}=\frac{2 x^{2}+2 x y+2 x z-2 y z}{2(x+y)(x+z)}=1-\frac{2 y z}{(x+y)(x+z)} .
$$

We use this result to apply the law of Cosines in triangle $A Z C$ :

$$
\begin{align*}
(p+q)^{2} & =x^{2}+(x+z)^{2}-2 x(x+z) \cos \alpha \\
& =x^{2}+(x+z)^{2}-2 x(x+z) \cdot\left(1-\frac{2 y z}{(x+y)(x+z)}\right) \\
& =z^{2}+\frac{4 x y z}{x+y} \tag{3}
\end{align*}
$$

Now in triangle $Z C S$ the segment $G S$ is a cevian, so with Stewart's theorem we have $p y^{2}+q(y+z)^{2}=(p+q)\left(G S^{2}+p q\right)$, hence

$$
G S^{2}=\frac{p}{p+q} \cdot y^{2}+\frac{q}{p+q} \cdot(y+z)^{2}-\frac{p}{p+q} \cdot \frac{q}{p+q} \cdot(p+q)^{2} .
$$

Replacing the $p$ 's and $q$ 's herein by (2) and (3) yields

$$
\begin{aligned}
G S^{2} & =\frac{x y}{N} y^{2}+\frac{y z+z x}{N}(y+z)^{2}-\frac{x y}{N} \cdot \frac{y z+z x}{N} \cdot\left(z^{2}+\frac{4 x y z}{x+y}\right) \\
& =\frac{x y^{3}}{N}+\underbrace{\frac{y z(y+z)^{2}}{N}}_{M_{1}}+\frac{z x(y+z)^{2}}{N}-\frac{x y z^{3}(x+y)}{N^{2}}-\underbrace{\frac{4 x^{2} y^{2} z^{2}}{N^{2}}}_{M_{2}} \\
& =\frac{x y^{3}+z x(y+z)^{2}}{N}-\frac{x y z^{3}(x+y)}{N^{2}}+M_{1}-M_{2} \\
& =\underbrace{\frac{x\left(y^{3}+y^{2} z+y z^{2}+z^{3}\right)}{N}+\frac{x y z^{2} N}{N^{2}}-\frac{x y z^{3}(x+y)}{N^{2}}+M_{1}-M_{2}}_{M_{3}} \\
& =\frac{x^{2} y^{2} z^{2}+x y^{2} z^{3}+x^{2} y z^{3}-x^{2} y z^{3}-x y^{2} z^{3}}{N^{2}}+M_{1}-M_{2}+M_{3} \\
& =\frac{x^{2} y^{2} z^{2}}{N^{2}}+M_{1}-M_{2}+M_{3},
\end{aligned}
$$

a term that is symmetric in $y$ and $z$, indeed.

Comment. $G$ is known as Gergonne's point of $\triangle A B C$.

## G4 UNK (United Kingdom)

Given a cyclic quadrilateral $A B C D$, let the diagonals $A C$ and $B D$ meet at $E$ and the lines $A D$ and $B C$ meet at $F$. The midpoints of $A B$ and $C D$ are $G$ and $H$, respectively. Show that $E F$ is tangent at $E$ to the circle through the points $E, G$, and $H$.

Solution 1. It suffices to show that $\angle H E F=\angle H G E$ (see Figure 1), since in circle $E G H$ the angle over the chord $E H$ at $G$ equals the angle between the tangent at $E$ and $E H$.
First, $\angle B A D=180^{\circ}-\angle D C B=\angle F C D$. Since triangles $F A B$ and $F C D$ have also a common interior angle at $F$, they are similar.


Figure 1
Denote by $\mathcal{T}$ the transformation consisting of a reflection at the bisector of $\angle D F C$ followed by a dilation with center $F$ and factor of $\frac{F A}{F C}$. Then $\mathcal{T}$ maps $F$ to $F, C$ to $A, D$ to $B$, and $H$ to $G$. To see this, note that $\triangle F C A \sim \triangle F D B$, so $\frac{F A}{F C}=\frac{F B}{F D}$. Moreover, as $\angle A D B=\angle A C B$, the image of the line $D E$ under $\mathcal{T}$ is parallel to $A C$ (and passes through $B$ ) and similarly the image of $C E$ is parallel to $D B$ and passes through $A$. Hence $E$ is mapped to the point $X$ which is the fourth vertex of the parallelogram $B E A X$. Thus, in particular $\angle H E F=\angle F X G$.
As $G$ is the midpoint of the diagonal $A B$ of the parallelogram $B E A X$, it is also the midpoint of $E X$. In particular, $E, G, X$ are collinear, and $E X=2 \cdot E G$.
Denote by $Y$ the fourth vertex of the parallelogram $D E C Y$. By an analogous reasoning as before, it follows that $\mathcal{T}$ maps $Y$ to $E$, thus $E, H, Y$ are collinear with $E Y=2 \cdot E H$. Therefore, by the intercept theorem, $H G \| X Y$.

From the construction of $\mathcal{T}$ it is clear that the lines $F X$ and $F E$ are symmetric with respect to the bisector of $\angle D F C$, as are $F Y$ and $F E$. Thus, $F, X, Y$ are collinear, which together with $H G \| X Y$ implies $\angle F X E=\angle H G E$. This completes the proof.

Solution 2. We use the following
Lemma (Gauß). Let $A B C D$ be a quadrilateral. Let $A B$ and $C D$ intersect at $P$, and $B C$ and $D A$ intersect at $Q$. Then the midpoints $K, L, M$ of $A C, B D$, and $P Q$, respectively, are collinear.
Proof: Let us consider the points $Z$ that fulfill the equation

$$
\begin{equation*}
(A B Z)+(C D Z)=(B C Z)+(D A Z) \tag{1}
\end{equation*}
$$

where $(R S T)$ denotes the oriented area of the triangle $R S T$ (see Figure 2).


Figure 2
As (1) is linear in $Z$, it can either characterize a line, or be contradictory, or be trivially fulfilled for all $Z$ in the plane. If (1) was fulfilled for all $Z$, then it would hold for $Z=A, Z=B$, which gives $(C D A)=(B C A),(C D B)=(D A B)$, respectively, i.e. the diagonals of $A B C D$ would bisect each other, thus $A B C D$ would be a parallelogram. This contradicts the hypothesis that $A D$ and $B C$ intersect. Since $E, F, G$ fulfill (1), it is the equation of a line which completes the proof of the lemma.
Now consider the parallelograms $E A X B$ and $E C Y D$ (see Figure 1). Then $G, H$ are the midpoints of $E X, E Y$, respectively. Let $M$ be the midpoint of $E F$. By applying the Lemma to the (re-entrant) quadrilateral $A D B C$, it is evident that $G, H$, and $M$ are collinear. A dilation by a factor of 2 with center $E$ shows that $X, Y, F$ are collinear. Since $A X \| D E$ and $B X \| C E$, we have pairwise equal interior angles in the quadrilaterals $F D E C$ and $F B X A$. Since we have also $\angle E B A=\angle D C A=\angle C D Y$, the quadrilaterals are similar. Thus, $\angle F X A=\angle C E F$.
Clearly the parallelograms $E C Y D$ and $E B X A$ are similar, too, thus $\angle E X A=\angle C E Y$. Consequently, $\angle F X E=\angle F X A-\angle E X A=\angle C E F-\angle C E Y=\angle Y E F$. By the converse of the tangent-chord angle theorem $E F$ is tangent to the circle $X E Y$. A dilation by a factor of $\frac{1}{2}$ completes the proof.

Solution 3. As in Solution 2, $G, H, M$ are proven to be collinear. It suffices to show that $M E^{2}=M G \cdot M H$. If $\boldsymbol{p}=\overrightarrow{O P}$ denotes the vector from circumcenter $O$ to point $P$, the claim becomes

$$
\left(\frac{\boldsymbol{e}-\boldsymbol{f}}{2}\right)^{2}=\left(\frac{\boldsymbol{e}+\boldsymbol{f}}{2}-\frac{\boldsymbol{a}+\boldsymbol{b}}{2}\right)\left(\frac{\boldsymbol{e}+\boldsymbol{f}}{2}-\frac{\boldsymbol{c}+\boldsymbol{d}}{2}\right)
$$

or equivalently

$$
\begin{equation*}
4 \boldsymbol{e} \boldsymbol{f}-(\boldsymbol{e}+\boldsymbol{f})(\boldsymbol{a}+\boldsymbol{b}+\boldsymbol{c}+\boldsymbol{d})+(\boldsymbol{a}+\boldsymbol{b})(\boldsymbol{c}+\boldsymbol{d})=0 \tag{2}
\end{equation*}
$$

With $R$ as the circumradius of $A B C D$, we obtain for the powers $\mathcal{P}(E)$ and $\mathcal{P}(F)$ of $E$ and $F$, respectively, with respect to the circumcircle

$$
\begin{aligned}
& \mathcal{P}(E)=(\boldsymbol{e}-\boldsymbol{a})(\boldsymbol{e}-\boldsymbol{c})=(\boldsymbol{e}-\boldsymbol{b})(\boldsymbol{e}-\boldsymbol{d})=\boldsymbol{e}^{2}-R^{2} \\
& \mathcal{P}(F)=(\boldsymbol{f}-\boldsymbol{a})(\boldsymbol{f}-\boldsymbol{d})=(\boldsymbol{f}-\boldsymbol{b})(\boldsymbol{f}-\boldsymbol{c})=\boldsymbol{f}^{2}-R^{2}
\end{aligned}
$$

hence

$$
\begin{align*}
& (\boldsymbol{e}-\boldsymbol{a})(\boldsymbol{e}-\boldsymbol{c})=\boldsymbol{e}^{2}-R^{2},  \tag{3}\\
& (\boldsymbol{e}-\boldsymbol{b})(\boldsymbol{e}-\boldsymbol{d})=\boldsymbol{e}^{2}-R^{2},  \tag{4}\\
& (\boldsymbol{f}-\boldsymbol{a})(\boldsymbol{f}-\boldsymbol{d})=\boldsymbol{f}^{2}-R^{2},  \tag{5}\\
& (\boldsymbol{f}-\boldsymbol{b})(\boldsymbol{f}-\boldsymbol{c})=\boldsymbol{f}^{2}-R^{2} . \tag{6}
\end{align*}
$$

Since $F$ lies on the polar to $E$ with respect to the circumcircle, we have

$$
\begin{equation*}
4 \boldsymbol{e} \boldsymbol{f}=4 R^{2} \tag{7}
\end{equation*}
$$

Adding up (3) to (7) yields (2), as desired.

## G5 POL (Poland)

Let $P$ be a polygon that is convex and symmetric to some point $O$. Prove that for some parallelogram $R$ satisfying $P \subset R$ we have

$$
\frac{|R|}{|P|} \leq \sqrt{2}
$$

where $|R|$ and $|P|$ denote the area of the sets $R$ and $P$, respectively.

Solution 1. We will construct two parallelograms $R_{1}$ and $R_{3}$, each of them containing $P$, and prove that at least one of the inequalities $\left|R_{1}\right| \leq \sqrt{2}|P|$ and $\left|R_{3}\right| \leq \sqrt{2}|P|$ holds (see Figure 1). First we will construct a parallelogram $R_{1} \supseteq P$ with the property that the midpoints of the sides of $R_{1}$ are points of the boundary of $P$.
Choose two points $A$ and $B$ of $P$ such that the triangle $O A B$ has maximal area. Let $a$ be the line through $A$ parallel to $O B$ and $b$ the line through $B$ parallel to $O A$. Let $A^{\prime}, B^{\prime}, a^{\prime}$ and $b^{\prime}$ be the points or lines, that are symmetric to $A, B, a$ and $b$, respectively, with respect to $O$. Now let $R_{1}$ be the parallelogram defined by $a, b, a^{\prime}$ and $b^{\prime}$.


Figure 1
Obviously, $A$ and $B$ are located on the boundary of the polygon $P$, and $A, B, A^{\prime}$ and $B^{\prime}$ are midpoints of the sides of $R_{1}$. We note that $P \subseteq R_{1}$. Otherwise, there would be a point $Z \in P$ but $Z \notin R_{1}$, i.e., one of the lines $a, b, a^{\prime}$ or $b^{\prime}$ were between $O$ and $Z$. If it is $a$, we have $|O Z B|>|O A B|$, which is contradictory to the choice of $A$ and $B$. If it is one of the lines $b, a^{\prime}$ or $b^{\prime}$ almost identical arguments lead to a similar contradiction.
Let $R_{2}$ be the parallelogram $A B A^{\prime} B^{\prime}$. Since $A$ and $B$ are points of $P$, segment $A B \subset P$ and so $R_{2} \subset R_{1}$. Since $A, B, A^{\prime}$ and $B^{\prime}$ are midpoints of the sides of $R_{1}$, an easy argument yields

$$
\begin{equation*}
\left|R_{1}\right|=2 \cdot\left|R_{2}\right| . \tag{1}
\end{equation*}
$$

Let $R_{3}$ be the smallest parallelogram enclosing $P$ defined by lines parallel to $A B$ and $B A^{\prime}$. Obviously $R_{2} \subset R_{3}$ and every side of $R_{3}$ contains at least one point of the boundary of $P$. Denote by $C$ the intersection point of $a$ and $b$, by $X$ the intersection point of $A B$ and $O C$, and by $X^{\prime}$ the intersection point of $X C$ and the boundary of $R_{3}$. In a similar way denote by $D$
the intersection point of $b$ and $a^{\prime}$, by $Y$ the intersection point of $A^{\prime} B$ and $O D$, and by $Y^{\prime}$ the intersection point of $Y D$ and the boundary of $R_{3}$.
Note that $O C=2 \cdot O X$ and $O D=2 \cdot O Y$, so there exist real numbers $x$ and $y$ with $1 \leq x, y \leq 2$ and $O X^{\prime}=x \cdot O X$ and $O Y^{\prime}=y \cdot O Y$. Corresponding sides of $R_{3}$ and $R_{2}$ are parallel which yields

$$
\begin{equation*}
\left|R_{3}\right|=x y \cdot\left|R_{2}\right| . \tag{2}
\end{equation*}
$$

The side of $R_{3}$ containing $X^{\prime}$ contains at least one point $X^{*}$ of $P$; due to the convexity of $P$ we have $A X^{*} B \subset P$. Since this side of the parallelogram $R_{3}$ is parallel to $A B$ we have $\left|A X^{*} B\right|=\left|A X^{\prime} B\right|$, so $\left|O A X^{\prime} B\right|$ does not exceed the area of $P$ confined to the sector defined by the rays $O B$ and $O A$. In a similar way we conclude that $\left|O B^{\prime} Y^{\prime} A^{\prime}\right|$ does not exceed the area of $P$ confined to the sector defined by the rays $O B$ and $O A^{\prime}$. Putting things together we have $\left|O A X^{\prime} B\right|=x \cdot|O A B|,\left|O B D A^{\prime}\right|=y \cdot\left|O B A^{\prime}\right|$. Since $|O A B|=\left|O B A^{\prime}\right|$, we conclude that $|P| \geq 2 \cdot\left|A X^{\prime} B Y^{\prime} A^{\prime}\right|=2 \cdot\left(x \cdot|O A B|+y \cdot\left|O B A^{\prime}\right|\right)=4 \cdot \frac{x+y}{2} \cdot|O A B|=\frac{x+y}{2} \cdot R_{2}$; this is in short

$$
\begin{equation*}
\frac{x+y}{2} \cdot\left|R_{2}\right| \leq|P| . \tag{3}
\end{equation*}
$$

Since all numbers concerned are positive, we can combine (1)-(3). Using the arithmetic-geometric-mean inequality we obtain

$$
\left|R_{1}\right| \cdot\left|R_{3}\right|=2 \cdot\left|R_{2}\right| \cdot x y \cdot\left|R_{2}\right| \leq 2 \cdot\left|R_{2}\right|^{2}\left(\frac{x+y}{2}\right)^{2} \leq 2 \cdot|P|^{2}
$$

This implies immediately the desired result $\left|R_{1}\right| \leq \sqrt{2} \cdot|P|$ or $\left|R_{3}\right| \leq \sqrt{2} \cdot|P|$.

Solution 2. We construct the parallelograms $R_{1}, R_{2}$ and $R_{3}$ in the same way as in Solution 1 and will show that $\frac{\left|R_{1}\right|}{|P|} \leq \sqrt{2}$ or $\frac{\left|R_{3}\right|}{|P|} \leq \sqrt{2}$.


Figure 2
Recall that affine one-to-one maps of the plane preserve the ratio of areas of subsets of the plane. On the other hand, every parallelogram can be transformed with an affine map onto a square. It follows that without loss of generality we may assume that $R_{1}$ is a square (see Figure 2).
Then $R_{2}$, whose vertices are the midpoints of the sides of $R_{1}$, is a square too, and $R_{3}$, whose sides are parallel to the diagonals of $R_{1}$, is a rectangle.
Let $a>0, b \geq 0$ and $c \geq 0$ be the distances introduced in Figure 2. Then $\left|R_{1}\right|=2 a^{2}$ and
$\left|R_{3}\right|=(a+2 b)(a+2 c)$.
Points $A, A^{\prime}, B$ and $B^{\prime}$ are in the convex polygon $P$. Hence the square $A B A^{\prime} B^{\prime}$ is a subset of $P$. Moreover, each of the sides of the rectangle $R_{3}$ contains a point of $P$, otherwise $R_{3}$ would not be minimal. It follows that

$$
|P| \geq a^{2}+2 \cdot \frac{a b}{2}+2 \cdot \frac{a c}{2}=a(a+b+c)
$$

Now assume that both $\frac{\left|R_{1}\right|}{|P|}>\sqrt{2}$ and $\frac{\left|R_{3}\right|}{|P|}>\sqrt{2}$, then

$$
2 a^{2}=\left|R_{1}\right|>\sqrt{2} \cdot|P| \geq \sqrt{2} \cdot a(a+b+c)
$$

and

$$
(a+2 b)(a+2 c)=\left|R_{3}\right|>\sqrt{2} \cdot|P| \geq \sqrt{2} \cdot a(a+b+c)
$$

All numbers concerned are positive, so after multiplying these inequalities we get

$$
2 a^{2}(a+2 b)(a+2 c)>2 a^{2}(a+b+c)^{2}
$$

But the arithmetic-geometric-mean inequality implies the contradictory result

$$
2 a^{2}(a+2 b)(a+2 c) \leq 2 a^{2}\left(\frac{(a+2 b)+(a+2 c)}{2}\right)^{2}=2 a^{2}(a+b+c)^{2}
$$

Hence $\frac{\left|R_{1}\right|}{|P|} \leq \sqrt{2}$ or $\frac{\left|R_{3}\right|}{|P|} \leq \sqrt{2}$, as desired.

## G6 UKR (Ukraine)

Let the sides $A D$ and $B C$ of the quadrilateral $A B C D$ (such that $A B$ is not parallel to $C D$ ) intersect at point $P$. Points $O_{1}$ and $O_{2}$ are the circumcenters and points $H_{1}$ and $H_{2}$ are the orthocenters of triangles $A B P$ and $D C P$, respectively. Denote the midpoints of segments $O_{1} H_{1}$ and $O_{2} H_{2}$ by $E_{1}$ and $E_{2}$, respectively. Prove that the perpendicular from $E_{1}$ on $C D$, the perpendicular from $E_{2}$ on $A B$ and the line $H_{1} H_{2}$ are concurrent.

Solution 1. We keep triangle $A B P$ fixed and move the line $C D$ parallel to itself uniformly, i.e. linearly dependent on a single parameter $\lambda$ (see Figure 1). Then the points $C$ and $D$ also move uniformly. Hence, the points $O_{2}, H_{2}$ and $E_{2}$ move uniformly, too. Therefore also the perpendicular from $E_{2}$ on $A B$ moves uniformly. Obviously, the points $O_{1}, H_{1}, E_{1}$ and the perpendicular from $E_{1}$ on $C D$ do not move at all. Hence, the intersection point $S$ of these two perpendiculars moves uniformly. Since $H_{1}$ does not move, while $H_{2}$ and $S$ move uniformly along parallel lines (both are perpendicular to $C D$ ), it is sufficient to prove their collinearity for two different positions of $C D$.


Figure 1
Let $C D$ pass through either point $A$ or point $B$. Note that by hypothesis these two cases are different. We will consider the case $A \in C D$, i.e. $A=D$. So we have to show that the perpendiculars from $E_{1}$ on $A C$ and from $E_{2}$ on $A B$ intersect on the altitude $A H$ of triangle $A B C$ (see Figure 2).


Figure 2

To this end, we consider the midpoints $A_{1}, B_{1}, C_{1}$ of $B C, C A, A B$, respectively. As $E_{1}$ is the center of Feuerbach's circle (nine-point circle) of $\triangle A B P$, we have $E_{1} C_{1}=E_{1} H$. Similarly, $E_{2} B_{1}=E_{2} H$. Note further that a point $X$ lies on the perpendicular from $E_{1}$ on $A_{1} C_{1}$ if and only if

$$
X C_{1}^{2}-X A_{1}^{2}=E_{1} C_{1}^{2}-E_{1} A_{1}^{2}
$$

Similarly, the perpendicular from $E_{2}$ on $A_{1} B_{1}$ is characterized by

$$
X A_{1}^{2}-X B_{1}^{2}=E_{2} A_{1}^{2}-E_{2} B_{1}^{2}
$$

The line $H_{1} H_{2}$, which is perpendicular to $B_{1} C_{1}$ and contains $A$, is given by

$$
X B_{1}^{2}-X C_{1}^{2}=A B_{1}^{2}-A C_{1}^{2}
$$

The three lines are concurrent if and only if

$$
\begin{aligned}
0 & =X C_{1}^{2}-X A_{1}^{2}+X A_{1}^{2}-X B_{1}^{2}+X B_{1}^{2}-X C_{1}^{2} \\
& =E_{1} C_{1}^{2}-E_{1} A_{1}^{2}+E_{2} A_{1}^{2}-E_{2} B_{1}^{2}+A B_{1}^{2}-A C_{1}^{2} \\
& =-E_{1} A_{1}^{2}+E_{2} A_{1}^{2}+E_{1} H^{2}-E_{2} H^{2}+A B_{1}^{2}-A C_{1}^{2}
\end{aligned}
$$

i.e. it suffices to show that

$$
E_{1} A_{1}^{2}-E_{2} A_{1}^{2}-E_{1} H^{2}+E_{2} H^{2}=\frac{A C^{2}-A B^{2}}{4}
$$

We have

$$
\frac{A C^{2}-A B^{2}}{4}=\frac{H C^{2}-H B^{2}}{4}=\frac{(H C+H B)(H C-H B)}{4}=\frac{H A_{1} \cdot B C}{2}
$$

Let $F_{1}, F_{2}$ be the projections of $E_{1}, E_{2}$ on $B C$. Obviously, these are the midpoints of $H P_{1}$,
$H P_{2}$, where $P_{1}, P_{2}$ are the midpoints of $P B$ and $P C$ respectively. Then

$$
\begin{aligned}
& E_{1} A_{1}^{2}-E_{2} A_{1}^{2}-E_{1} H^{2}+E_{2} H^{2} \\
& =F_{1} A_{1}^{2}-F_{1} H^{2}-F_{2} A_{1}^{2}+F_{2} H^{2} \\
& =\left(F_{1} A_{1}-F_{1} H\right)\left(F_{1} A_{1}+F_{1} H\right)-\left(F_{2} A_{1}-F_{2} H\right)\left(F_{2} A_{1}+F_{2} H\right) \\
& =A_{1} H \cdot\left(A_{1} P_{1}-A_{1} P_{2}\right) \\
& =\frac{A_{1} H \cdot B C}{2} \\
& =\frac{A C^{2}-A B^{2}}{4}
\end{aligned}
$$

which proves the claim.

Solution 2. Let the perpendicular from $E_{1}$ on $C D$ meet $P H_{1}$ at $X$, and the perpendicular from $E_{2}$ on $A B$ meet $P H_{2}$ at $Y$ (see Figure 3). Let $\varphi$ be the intersection angle of $A B$ and $C D$. Denote by $M, N$ the midpoints of $P H_{1}, P H_{2}$ respectively.


Figure 3
We will prove now that triangles $E_{1} X M$ and $E_{2} Y N$ have equal angles at $E_{1}, E_{2}$, and supplementary angles at $X, Y$.

In the following, angles are understood as oriented, and equalities of angles modulo $180^{\circ}$.
Let $\alpha=\angle H_{2} P D, \psi=\angle D P C, \beta=\angle C P H_{1}$. Then $\alpha+\psi+\beta=\varphi, \angle E_{1} X H_{1}=\angle H_{2} Y E_{2}=\varphi$, thus $\angle M X E_{1}+\angle N Y E_{2}=180^{\circ}$.
By considering the Feuerbach circle of $\triangle A B P$ whose center is $E_{1}$ and which goes through $M$, we have $\angle E_{1} M H_{1}=\psi+2 \beta$. Analogous considerations with the Feuerbach circle of $\triangle D C P$ yield $\angle H_{2} N E_{2}=\psi+2 \alpha$. Hence indeed $\angle X E_{1} M=\varphi-(\psi+2 \beta)=(\psi+2 \alpha)-\varphi=\angle Y E_{2} N$. It follows now that

$$
\frac{X M}{M E_{1}}=\frac{Y N}{N E_{2}} .
$$

Furthermore, $M E_{1}$ is half the circumradius of $\triangle A B P$, while $P H_{1}$ is the distance of $P$ to the orthocenter of that triangle, which is twice the circumradius times the cosine of $\psi$. Together
with analogous reasoning for $\triangle D C P$ we have

$$
\frac{M E_{1}}{P H_{1}}=\frac{1}{4 \cos \psi}=\frac{N E_{2}}{P H_{2}} .
$$

By multiplication,

$$
\frac{X M}{P H_{1}}=\frac{Y N}{P H_{2}}
$$

and therefore

$$
\frac{P X}{X H_{1}}=\frac{H_{2} Y}{Y P}
$$

Let $E_{1} X, E_{2} Y$ meet $H_{1} H_{2}$ in $R, S$ respectively.
Applying the intercept theorem to the parallels $E_{1} X, P H_{2}$ and center $H_{1}$ gives

$$
\frac{H_{2} R}{R H_{1}}=\frac{P X}{X H_{1}},
$$

while with parallels $E_{2} Y, P H_{1}$ and center $H_{2}$ we obtain

$$
\frac{H_{2} S}{S H_{1}}=\frac{H_{2} Y}{Y P}
$$

Combination of the last three equalities yields that $R$ and $S$ coincide.

## G7 IRN (Islamic Republic of Iran)

Let $A B C$ be a triangle with incenter $I$ and let $X, Y$ and $Z$ be the incenters of the triangles $B I C, C I A$ and $A I B$, respectively. Let the triangle $X Y Z$ be equilateral. Prove that $A B C$ is equilateral too.

Solution. $A Z, A I$ and $A Y$ divide $\angle B A C$ into four equal angles; denote them by $\alpha$. In the same way we have four equal angles $\beta$ at $B$ and four equal angles $\gamma$ at $C$. Obviously $\alpha+\beta+\gamma=\frac{180^{\circ}}{4}=45^{\circ} ;$ and $0^{\circ}<\alpha, \beta, \gamma<45^{\circ}$.


Easy calculations in various triangles yield $\angle B I C=180^{\circ}-2 \beta-2 \gamma=180^{\circ}-\left(90^{\circ}-2 \alpha\right)=$ $90^{\circ}+2 \alpha$, hence (for $X$ is the incenter of triangle $B C I$, so $I X$ bisects $\angle B I C$ ) we have $\angle X I C=$ $\angle B I X=\frac{1}{2} \angle B I C=45^{\circ}+\alpha$ and with similar aguments $\angle C I Y=\angle Y I A=45^{\circ}+\beta$ and $\angle A I Z=\angle Z I B=45^{\circ}+\gamma$. Furthermore, we have $\angle X I Y=\angle X I C+\angle C I Y=\left(45^{\circ}+\alpha\right)+$ $\left(45^{\circ}+\beta\right)=135^{\circ}-\gamma, \angle Y I Z=135^{\circ}-\alpha$, and $\angle Z I X=135^{\circ}-\beta$.

Now we calculate the lengths of $I X, I Y$ and $I Z$ in terms of $\alpha, \beta$ and $\gamma$. The perpendicular from $I$ on $C X$ has length $I X \cdot \sin \angle C X I=I X \cdot \sin \left(90^{\circ}+\beta\right)=I X \cdot \cos \beta$. But $C I$ bisects $\angle Y C X$, so the perpendicular from $I$ on $C Y$ has the same length, and we conclude

$$
I X \cdot \cos \beta=I Y \cdot \cos \alpha
$$

To make calculations easier we choose a length unit that makes $I X=\cos \alpha$. Then $I Y=\cos \beta$ and with similar arguments $I Z=\cos \gamma$.
Since $X Y Z$ is equilateral we have $Z X=Z Y$. The law of Cosines in triangles $X Y I, Y Z I$ yields

$$
\begin{aligned}
& Z X^{2}=Z Y^{2} \\
\Longrightarrow & I Z^{2}+I X^{2}-2 \cdot I Z \cdot I X \cdot \cos \angle Z I X=I Z^{2}+I Y^{2}-2 \cdot I Z \cdot I Y \cdot \cos \angle Y I Z \\
\Longrightarrow & I X^{2}-I Y^{2}=2 \cdot I Z \cdot(I X \cdot \cos \angle Z I X-I Y \cdot \cos \angle Y I Z) \\
\Longrightarrow & \underbrace{\cos ^{2} \alpha-\cos ^{2} \beta}_{\text {L.H.S. }}=\underbrace{2 \cdot \cos \gamma \cdot\left(\cos \alpha \cdot \cos \left(135^{\circ}-\beta\right)-\cos \beta \cdot \cos \left(135^{\circ}-\alpha\right)\right)}_{\text {R.H.S. }} .
\end{aligned}
$$

A transformation of the left-hand side (L.H.S.) yields

$$
\begin{aligned}
\text { L.H.S. } & =\cos ^{2} \alpha \cdot\left(\sin ^{2} \beta+\cos ^{2} \beta\right)-\cos ^{2} \beta \cdot\left(\sin ^{2} \alpha+\cos ^{2} \alpha\right) \\
& =\cos ^{2} \alpha \cdot \sin ^{2} \beta-\cos ^{2} \beta \cdot \sin ^{2} \alpha
\end{aligned}
$$

$$
\begin{aligned}
& =(\cos \alpha \cdot \sin \beta+\cos \beta \cdot \sin \alpha) \cdot(\cos \alpha \cdot \sin \beta-\cos \beta \cdot \sin \alpha) \\
& =\sin (\beta+\alpha) \cdot \sin (\beta-\alpha)=\sin \left(45^{\circ}-\gamma\right) \cdot \sin (\beta-\alpha)
\end{aligned}
$$

whereas a transformation of the right-hand side (R.H.S.) leads to

$$
\begin{aligned}
\text { R.H.S. } & =2 \cdot \cos \gamma \cdot\left(\cos \alpha \cdot\left(-\cos \left(45^{\circ}+\beta\right)\right)-\cos \beta \cdot\left(-\cos \left(45^{\circ}+\alpha\right)\right)\right) \\
& =2 \cdot \frac{\sqrt{2}}{2} \cdot \cos \gamma \cdot(\cos \alpha \cdot(\sin \beta-\cos \beta)+\cos \beta \cdot(\cos \alpha-\sin \alpha)) \\
& =\sqrt{2} \cdot \cos \gamma \cdot(\cos \alpha \cdot \sin \beta-\cos \beta \cdot \sin \alpha) \\
& =\sqrt{2} \cdot \cos \gamma \cdot \sin (\beta-\alpha) .
\end{aligned}
$$

Equating L.H.S. and R.H.S. we obtain

$$
\begin{aligned}
& \sin \left(45^{\circ}-\gamma\right) \cdot \sin (\beta-\alpha)=\sqrt{2} \cdot \cos \gamma \cdot \sin (\beta-\alpha) \\
\Longrightarrow & \sin (\beta-\alpha) \cdot\left(\sqrt{2} \cdot \cos \gamma-\sin \left(45^{\circ}-\gamma\right)\right)=0 \\
\Longrightarrow & \alpha=\beta \text { or } \sqrt{2} \cdot \cos \gamma=\sin \left(45^{\circ}-\gamma\right) .
\end{aligned}
$$

But $\gamma<45^{\circ}$; so $\sqrt{2} \cdot \cos \gamma>\cos \gamma>\cos 45^{\circ}=\sin 45^{\circ}>\sin \left(45^{\circ}-\gamma\right)$. This leaves $\alpha=\beta$. With similar reasoning we have $\alpha=\gamma$, which means triangle $A B C$ must be equilateral.

## G8 BGR (Bulgaria)

Let $A B C D$ be a circumscribed quadrilateral. Let $g$ be a line through $A$ which meets the segment $B C$ in $M$ and the line $C D$ in $N$. Denote by $I_{1}, I_{2}$, and $I_{3}$ the incenters of $\triangle A B M$, $\triangle M N C$, and $\triangle N D A$, respectively. Show that the orthocenter of $\triangle I_{1} I_{2} I_{3}$ lies on $g$.

Solution 1. Let $k_{1}, k_{2}$ and $k_{3}$ be the incircles of triangles $A B M, M N C$, and $N D A$, respectively (see Figure 1). We shall show that the tangent $h$ from $C$ to $k_{1}$ which is different from $C B$ is also tangent to $k_{3}$.


Figure 1
To this end, let $X$ denote the point of intersection of $g$ and $h$. Then $A B C X$ and $A B C D$ are circumscribed quadrilaterals, whence

$$
C D-C X=(A B+C D)-(A B+C X)=(B C+A D)-(B C+A X)=A D-A X
$$

i.e.

$$
A X+C D=C X+A D
$$

which in turn reveals that the quadrilateral $A X C D$ is also circumscribed. Thus $h$ touches indeed the circle $k_{3}$.
Moreover, we find that $\angle I_{3} C I_{1}=\angle I_{3} C X+\angle X C I_{1}=\frac{1}{2}(\angle D C X+\angle X C B)=\frac{1}{2} \angle D C B=$ $\frac{1}{2}\left(180^{\circ}-\angle M C N\right)=180^{\circ}-\angle M I_{2} N=\angle I_{3} I_{2} I_{1}$, from which we conclude that $C, I_{1}, I_{2}, I_{3}$ are concyclic.
Let now $L_{1}$ and $L_{3}$ be the reflection points of $C$ with respect to the lines $I_{2} I_{3}$ and $I_{1} I_{2}$ respectively. Since $I_{1} I_{2}$ is the angle bisector of $\angle N M C$, it follows that $L_{3}$ lies on $g$. By analogous reasoning, $L_{1}$ lies on $g$.
Let $H$ be the orthocenter of $\triangle I_{1} I_{2} I_{3}$. We have $\angle I_{2} L_{3} I_{1}=\angle I_{1} C I_{2}=\angle I_{1} I_{3} I_{2}=180^{\circ}-\angle I_{1} H I_{2}$, which entails that the quadrilateral $I_{2} H I_{1} L_{3}$ is cyclic. Analogously, $I_{3} H L_{1} I_{2}$ is cyclic.

Then, working with oriented angles modulo $180^{\circ}$, we have

$$
\angle L_{3} H I_{2}=\angle L_{3} I_{1} I_{2}=\angle I_{2} I_{1} C=\angle I_{2} I_{3} C=\angle L_{1} I_{3} I_{2}=\angle L_{1} H I_{2},
$$

whence $L_{1}, L_{3}$, and $H$ are collinear. By $L_{1} \neq L_{3}$, the claim follows.

Comment. The last part of the argument essentially reproves the following fact: The Simson line of a point $P$ lying on the circumcircle of a triangle $A B C$ with respect to that triangle bisects the line segment connecting $P$ with the orthocenter of $A B C$.

Solution 2. We start by proving that $C, I_{1}, I_{2}$, and $I_{3}$ are concyclic.


Figure 2
To this end, notice first that $I_{2}, M, I_{1}$ are collinear, as are $N, I_{2}, I_{3}$ (see Figure 2). Denote by $\alpha, \beta, \gamma, \delta$ the internal angles of $A B C D$. By considerations in triangle $C M N$, it follows that $\angle I_{3} I_{2} I_{1}=\frac{\gamma}{2}$. We will show that $\angle I_{3} C I_{1}=\frac{\gamma}{2}$, too. Denote by $I$ the incenter of $A B C D$. Clearly, $I_{1} \in B I, I_{3} \in D I, \angle I_{1} A I_{3}=\frac{\alpha}{2}$.
Using the abbreviation $[X, Y Z]$ for the distance from point $X$ to the line $Y Z$, we have because of $\angle B A I_{1}=\angle I A I_{3}$ and $\angle I_{1} A I=\angle I_{3} A D$ that

$$
\frac{\left[I_{1}, A B\right]}{\left[I_{1}, A I\right]}=\frac{\left[I_{3}, A I\right]}{\left[I_{3}, A D\right]}
$$

Furthermore, consideration of the angle sums in $A I B, B I C, C I D$ and $D I A$ implies $\angle A I B+$ $\angle C I D=\angle B I C+\angle D I A=180^{\circ}$, from which we see

$$
\frac{\left[I_{1}, A I\right]}{\left[I_{3}, C I\right]}=\frac{I_{1} I}{I_{3} I}=\frac{\left[I_{1}, C I\right]}{\left[I_{3}, A I\right]} .
$$

Because of $\left[I_{1}, A B\right]=\left[I_{1}, B C\right],\left[I_{3}, A D\right]=\left[I_{3}, C D\right]$, multiplication yields

$$
\frac{\left[I_{1}, B C\right]}{\left[I_{3}, C I\right]}=\frac{\left[I_{1}, C I\right]}{\left[I_{3}, C D\right]}
$$

By $\angle D C I=\angle I C B=\gamma / 2$ it follows that $\angle I_{1} C B=\angle I_{3} C I$ which concludes the proof of the
above statement.
Let the perpendicular from $I_{1}$ on $I_{2} I_{3}$ intersect $g$ at $Z$. Then $\angle M I_{1} Z=90^{\circ}-\angle I_{3} I_{2} I_{1}=$ $90^{\circ}-\gamma / 2=\angle M C I_{2}$. Since we have also $\angle Z M I_{1}=\angle I_{2} M C$, triangles $M Z I_{1}$ and $M I_{2} C$ are similar. From this one easily proves that also $M I_{2} Z$ and $M C I_{1}$ are similar. Because $C, I_{1}, I_{2}$, and $I_{3}$ are concyclic, $\angle M Z I_{2}=\angle M I_{1} C=\angle N I_{3} C$, thus $N I_{2} Z$ and $N C I_{3}$ are similar, hence $N C I_{2}$ and $N I_{3} Z$ are similar. We conclude $\angle Z I_{3} I_{2}=\angle I_{2} C N=90^{\circ}-\gamma / 2$, hence $I_{1} I_{2} \perp Z I_{3}$. This completes the proof.

## Number Theory

## N1 AUS (Australia)

A social club has $n$ members. They have the membership numbers $1,2, \ldots, n$, respectively. From time to time members send presents to other members, including items they have already received as presents from other members. In order to avoid the embarrassing situation that a member might receive a present that he or she has sent to other members, the club adds the following rule to its statutes at one of its annual general meetings:
"A member with membership number $a$ is permitted to send a present to a member with membership number $b$ if and only if $a(b-1)$ is a multiple of $n$."
Prove that, if each member follows this rule, none will receive a present from another member that he or she has already sent to other members.

Alternative formulation: Let $G$ be a directed graph with $n$ vertices $v_{1}, v_{2}, \ldots, v_{n}$, such that there is an edge going from $v_{a}$ to $v_{b}$ if and only if $a$ and $b$ are distinct and $a(b-1)$ is a multiple of $n$. Prove that this graph does not contain a directed cycle.

Solution 1. Suppose there is an edge from $v_{i}$ to $v_{j}$. Then $i(j-1)=i j-i=k n$ for some integer $k$, which implies $i=i j-k n$. If $\operatorname{gcd}(i, n)=d$ and $\operatorname{gcd}(j, n)=e$, then $e$ divides $i j-k n=i$ and thus $e$ also divides $d$. Hence, if there is an edge from $v_{i}$ to $v_{j}$, then $\operatorname{gcd}(j, n) \mid \operatorname{gcd}(i, n)$.
If there is a cycle in $G$, say $v_{i_{1}} \rightarrow v_{i_{2}} \rightarrow \cdots \rightarrow v_{i_{r}} \rightarrow v_{i_{1}}$, then we have

$$
\operatorname{gcd}\left(i_{1}, n\right)\left|\operatorname{gcd}\left(i_{r}, n\right)\right| \operatorname{gcd}\left(i_{r-1}, n\right)|\ldots| \operatorname{gcd}\left(i_{2}, n\right) \mid \operatorname{gcd}\left(i_{1}, n\right),
$$

which implies that all these greatest common divisors must be equal, say be equal to $t$.
Now we pick any of the $i_{k}$, without loss of generality let it be $i_{1}$. Then $i_{r}\left(i_{1}-1\right)$ is a multiple of $n$ and hence also (by dividing by $t$ ), $i_{1}-1$ is a multiple of $\frac{n}{t}$. Since $i_{1}$ and $i_{1}-1$ are relatively prime, also $t$ and $\frac{n}{t}$ are relatively prime. So, by the Chinese remainder theorem, the value of $i_{1}$ is uniquely determined modulo $n=t \cdot \frac{n}{t}$ by the value of $t$. But, as $i_{1}$ was chosen arbitrarily among the $i_{k}$, this implies that all the $i_{k}$ have to be equal, a contradiction.

Solution 2. If $a, b, c$ are integers such that $a b-a$ and $b c-b$ are multiples of $n$, then also $a c-a=a(b c-b)+(a b-a)-(a b-a) c$ is a multiple of $n$. This implies that if there is an edge from $v_{a}$ to $v_{b}$ and an edge from $v_{b}$ to $v_{c}$, then there also must be an edge from $v_{a}$ to $v_{c}$. Therefore, if there are any cycles at all, the smallest cycle must have length 2. But suppose the vertices $v_{a}$ and $v_{b}$ form such a cycle, i. e., $a b-a$ and $a b-b$ are both multiples of $n$. Then $a-b$ is also a multiple of $n$, which can only happen if $a=b$, which is impossible.

Solution 3. Suppose there was a cycle $v_{i_{1}} \rightarrow v_{i_{2}} \rightarrow \cdots \rightarrow v_{i_{r}} \rightarrow v_{i_{1}}$. Then $i_{1}\left(i_{2}-1\right)$ is a multiple of $n$, i.e., $i_{1} \equiv i_{1} i_{2} \bmod n$. Continuing in this manner, we get $i_{1} \equiv i_{1} i_{2} \equiv$ $i_{1} i_{2} i_{3} \equiv i_{1} i_{2} i_{3} \ldots i_{r} \bmod n$. But the same holds for all $i_{k}$, i. e., $i_{k} \equiv i_{1} i_{2} i_{3} \ldots i_{r} \bmod n$. Hence $i_{1} \equiv i_{2} \equiv \cdots \equiv i_{r} \bmod n$, which means $i_{1}=i_{2}=\cdots=i_{r}$, a contradiction.

Solution 4. Let $n=k$ be the smallest value of $n$ for which the corresponding graph has a cycle. We show that $k$ is a prime power.
If $k$ is not a prime power, it can be written as a product $k=d e$ of relatively prime integers greater than 1. Reducing all the numbers modulo $d$ yields a single vertex or a cycle in the corresponding graph on $d$ vertices, because if $a(b-1) \equiv 0 \bmod k$ then this equation also holds modulo $d$. But since the graph on $d$ vertices has no cycles, by the minimality of $k$, we must have that all the indices of the cycle are congruent modulo $d$. The same holds modulo $e$ and hence also modulo $k=d e$. But then all the indices are equal, which is a contradiction.
Thus $k$ must be a prime power $k=p^{m}$. There are no edges ending at $v_{k}$, so $v_{k}$ is not contained in any cycle. All edges not starting at $v_{k}$ end at a vertex belonging to a non-multiple of $p$, and all edges starting at a non-multiple of $p$ must end at $v_{1}$. But there is no edge starting at $v_{1}$. Hence there is no cycle.

Solution 5. Suppose there was a cycle $v_{i_{1}} \rightarrow v_{i_{2}} \rightarrow \cdots \rightarrow v_{i_{r}} \rightarrow v_{i_{1}}$. Let $q=p^{m}$ be a prime power dividing $n$. We claim that either $i_{1} \equiv i_{2} \equiv \cdots \equiv i_{r} \equiv 0 \bmod q$ or $i_{1} \equiv i_{2} \equiv \cdots \equiv i_{r} \equiv$ $1 \bmod q$.

Suppose that there is an $i_{s}$ not divisible by $q$. Then, as $i_{s}\left(i_{s+1}-1\right)$ is a multiple of $q, i_{s+1} \equiv$ $1 \bmod p$. Similarly, we conclude $i_{s+2} \equiv 1 \bmod p$ and so on. So none of the labels is divisible by $p$, but since $i_{s}\left(i_{s+1}-1\right)$ is a multiple of $q=p^{m}$ for all $s$, all $i_{s+1}$ are congruent to 1 modulo $q$. This proves the claim.
Now, as all the labels are congruent modulo all the prime powers dividing $n$, they must all be equal by the Chinese remainder theorem. This is a contradiction.

## N2 PER (Peru)

A positive integer $N$ is called balanced, if $N=1$ or if $N$ can be written as a product of an even number of not necessarily distinct primes. Given positive integers $a$ and $b$, consider the polynomial $P$ defined by $P(x)=(x+a)(x+b)$.
(a) Prove that there exist distinct positive integers $a$ and $b$ such that all the numbers $P(1), P(2)$, $\ldots, P(50)$ are balanced.
(b) Prove that if $P(n)$ is balanced for all positive integers $n$, then $a=b$.

Solution. Define a function $f$ on the set of positive integers by $f(n)=0$ if $n$ is balanced and $f(n)=1$ otherwise. Clearly, $f(n m) \equiv f(n)+f(m) \bmod 2$ for all positive integers $n, m$.
(a) Now for each positive integer $n$ consider the binary sequence $(f(n+1), f(n+2), \ldots, f(n+$ $50)$ ). As there are only $2^{50}$ different such sequences there are two different positive integers $a$ and $b$ such that

$$
(f(a+1), f(a+2), \ldots, f(a+50))=(f(b+1), f(b+2), \ldots, f(b+50))
$$

But this implies that for the polynomial $P(x)=(x+a)(x+b)$ all the numbers $P(1), P(2)$, $\ldots, P(50)$ are balanced, since for all $1 \leq k \leq 50$ we have $f(P(k)) \equiv f(a+k)+f(b+k) \equiv$ $2 f(a+k) \equiv 0 \bmod 2$.
(b) Now suppose $P(n)$ is balanced for all positive integers $n$ and $a<b$. Set $n=k(b-a)-a$ for sufficiently large $k$, such that $n$ is positive. Then $P(n)=k(k+1)(b-a)^{2}$, and this number can only be balanced, if $f(k)=f(k+1)$ holds. Thus, the sequence $f(k)$ must become constant for sufficiently large $k$. But this is not possible, as for every prime $p$ we have $f(p)=1$ and for every square $t^{2}$ we have $f\left(t^{2}\right)=0$.
Hence $a=b$.

Comment. Given a positive integer $k$, a computer search for the pairs of positive integers $(a, b)$, for which $P(1), P(2), \ldots, P(k)$ are all balanced yields the following results with minimal sum $a+b$ and $a<b$ :

| $k$ | 3 | 4 | 5 | 10 | 20 |
| :--- | :---: | :---: | :---: | :---: | :---: |
| $(a, b)$ | $(2,4)$ | $(6,11)$ | $(8,14)$ | $(20,34)$ | $(1751,3121)$ |

Therefore, trying to find $a$ and $b$ in part (a) of the problem cannot be done by elementary calculations.

## N3 EST (Estonia)

Let $f$ be a non-constant function from the set of positive integers into the set of positive integers, such that $a-b$ divides $f(a)-f(b)$ for all distinct positive integers $a, b$. Prove that there exist infinitely many primes $p$ such that $p$ divides $f(c)$ for some positive integer $c$.

Solution 1. Denote by $v_{p}(a)$ the exponent of the prime $p$ in the prime decomposition of $a$.
Assume that there are only finitely many primes $p_{1}, p_{2}, \ldots, p_{m}$ that divide some function value produced of $f$.
There are infinitely many positive integers $a$ such that $v_{p_{i}}(a)>v_{p_{i}}(f(1))$ for all $i=1,2, \ldots, m$, e.g. $a=\left(p_{1} p_{2} \ldots p_{m}\right)^{\alpha}$ with $\alpha$ sufficiently large. Pick any such $a$. The condition of the problem then yields $a \mid(f(a+1)-f(1))$. Assume $f(a+1) \neq f(1)$. Then we must have $v_{p_{i}}(f(a+1)) \neq$ $v_{p_{i}}(f(1))$ for at least one $i$. This yields $v_{p_{i}}(f(a+1)-f(1))=\min \left\{v_{p_{i}}(f(a+1)), v_{p_{i}}(f(1))\right\} \leq$ $v_{p_{1}}(f(1))<v_{p_{i}}(a)$. But this contradicts the fact that $a \mid(f(a+1)-f(1))$.
Hence we must have $f(a+1)=f(1)$ for all such $a$.
Now, for any positive integer $b$ and all such $a$, we have $(a+1-b) \mid(f(a+1)-f(b))$, i.e., $(a+1-b) \mid(f(1)-f(b))$. Since this is true for infinitely many positive integers $a$ we must have $f(b)=f(1)$. Hence $f$ is a constant function, a contradiction. Therefore, our initial assumption was false and there are indeed infinitely many primes $p$ dividing $f(c)$ for some positive integer c.

Solution 2. Assume that there are only finitely many primes $p_{1}, p_{2}, \ldots, p_{m}$ that divide some function value of $f$. Since $f$ is not identically 1 , we must have $m \geq 1$.
Then there exist non-negative integers $\alpha_{1}, \ldots, \alpha_{m}$ such that

$$
f(1)=p_{1}^{\alpha_{1}} p_{2}^{\alpha_{2}} \ldots p_{m}^{\alpha_{m}}
$$

We can pick a positive integer $r$ such that $f(r) \neq f(1)$. Let

$$
M=1+p_{1}^{\alpha_{1}+1} p_{2}^{\alpha_{2}+1} \ldots p_{m}^{\alpha_{m}+1} \cdot(f(r)+r)
$$

Then for all $i \in\{1, \ldots, m\}$ we have that $p_{i}^{\alpha_{i}+1}$ divides $M-1$ and hence by the condition of the problem also $f(M)-f(1)$. This implies that $f(M)$ is divisible by $p_{i}^{\alpha_{i}}$ but not by $p_{i}^{\alpha_{i}+1}$ for all $i$ and therefore $f(M)=f(1)$.
Hence

$$
\begin{aligned}
M-r & >p_{1}^{\alpha_{1}+1} p_{2}^{\alpha_{2}+1} \ldots p_{m}^{\alpha_{m}+1} \cdot(f(r)+r)-r \\
& \geq p_{1}^{\alpha_{1}+1} p_{2}^{\alpha_{2}+1} \ldots p_{m}^{\alpha_{m}+1}+(f(r)+r)-r \\
& >p_{1}^{\alpha_{1}} p_{2}^{\alpha_{2}} \ldots p_{m}^{\alpha_{m}}+f(r) \\
& \geq|f(M)-f(r)| .
\end{aligned}
$$

But since $M-r$ divides $f(M)-f(r)$ this can only be true if $f(r)=f(M)=f(1)$, which contradicts the choice of $r$.

Comment. In the case that $f$ is a polynomial with integer coefficients the result is well-known, see e.g. W. Schwarz, Einführung in die Methoden der Primzahltheorie, 1969.

## N4 PRK (Democratic People's Republic of Korea)

Find all positive integers $n$ such that there exists a sequence of positive integers $a_{1}, a_{2}, \ldots, a_{n}$ satisfying

$$
a_{k+1}=\frac{a_{k}^{2}+1}{a_{k-1}+1}-1
$$

for every $k$ with $2 \leq k \leq n-1$.

Solution 1. Such a sequence exists for $n=1,2,3,4$ and no other $n$. Since the existence of such a sequence for some $n$ implies the existence of such a sequence for all smaller $n$, it suffices to prove that $n=5$ is not possible and $n=4$ is possible.
Assume first that for $n=5$ there exists a sequence of positive integers $a_{1}, a_{2}, \ldots, a_{5}$ satisfying the conditions

$$
\begin{aligned}
& a_{2}^{2}+1=\left(a_{1}+1\right)\left(a_{3}+1\right), \\
& a_{3}^{2}+1=\left(a_{2}+1\right)\left(a_{4}+1\right), \\
& a_{4}^{2}+1=\left(a_{3}+1\right)\left(a_{5}+1\right) .
\end{aligned}
$$

Assume $a_{1}$ is odd, then $a_{2}$ has to be odd as well and as then $a_{2}^{2}+1 \equiv 2 \bmod 4, a_{3}$ has to be even. But this is a contradiction, since then the even number $a_{2}+1$ cannot divide the odd number $a_{3}^{2}+1$.
Hence $a_{1}$ is even.
If $a_{2}$ is odd, $a_{3}^{2}+1$ is even (as a multiple of $a_{2}+1$ ) and hence $a_{3}$ is odd, too. Similarly we must have $a_{4}$ odd as well. But then $a_{3}^{2}+1$ is a product of two even numbers $\left(a_{2}+1\right)\left(a_{4}+1\right)$ and thus is divisible by 4 , which is a contradiction as for odd $a_{3}$ we have $a_{3}^{2}+1 \equiv 2 \bmod 4$.
Hence $a_{2}$ is even. Furthermore $a_{3}+1$ divides the odd number $a_{2}^{2}+1$ and so $a_{3}$ is even. Similarly, $a_{4}$ and $a_{5}$ are even as well.
Now set $x=a_{2}$ and $y=a_{3}$. From the given condition we get $(x+1) \mid\left(y^{2}+1\right)$ and $(y+1) \mid\left(x^{2}+1\right)$. We will prove that there is no pair of positive even numbers $(x, y)$ satisfying these two conditions, thus yielding a contradiction to the assumption.
Assume there exists a pair $\left(x_{0}, y_{0}\right)$ of positive even numbers satisfying the two conditions $\left(x_{0}+1\right) \mid\left(y_{0}^{2}+1\right)$ and $\left(y_{0}+1\right) \mid\left(x_{0}^{2}+1\right)$.
Then one has $\left(x_{0}+1\right) \mid\left(y_{0}^{2}+1+x_{0}^{2}-1\right)$, i.e., $\left(x_{0}+1\right) \mid\left(x_{0}^{2}+y_{0}^{2}\right)$, and similarly $\left(y_{0}+1\right) \mid\left(x_{0}^{2}+y_{0}^{2}\right)$. Any common divisor $d$ of $x_{0}+1$ and $y_{0}+1$ must hence also divide the number $\left(x_{0}^{2}+1\right)+\left(y_{0}^{2}+1\right)-\left(x_{0}^{2}+y_{0}^{2}\right)=2$. But as $x_{0}+1$ and $y_{0}+1$ are both odd, we must have $d=1$. Thus $x_{0}+1$ and $y_{0}+1$ are relatively prime and therefore there exists a positive integer $k$ such that

$$
k(x+1)(y+1)=x^{2}+y^{2}
$$

has the solution $\left(x_{0}, y_{0}\right)$. We will show that the latter equation has no solution $(x, y)$ in positive even numbers.

Assume there is a solution. Pick the solution $\left(x_{1}, y_{1}\right)$ with the smallest sum $x_{1}+y_{1}$ and assume $x_{1} \geq y_{1}$. Then $x_{1}$ is a solution to the quadratic equation

$$
x^{2}-k\left(y_{1}+1\right) x+y_{1}^{2}-k\left(y_{1}+1\right)=0 .
$$

Let $x_{2}$ be the second solution, which by Vieta's theorem fulfills $x_{1}+x_{2}=k\left(y_{1}+1\right)$ and $x_{1} x_{2}=y_{1}^{2}-k\left(y_{1}+1\right)$. If $x_{2}=0$, the second equation implies $y_{1}^{2}=k\left(y_{1}+1\right)$, which is impossible, as $y_{1}+1>1$ cannot divide the relatively prime number $y_{1}^{2}$. Therefore $x_{2} \neq 0$.
Also we get $\left(x_{1}+1\right)\left(x_{2}+1\right)=x_{1} x_{2}+x_{1}+x_{2}+1=y_{1}^{2}+1$ which is odd, and hence $x_{2}$ must be even and positive. Also we have $x_{2}+1=\frac{y_{1}^{2}+1}{x_{1}+1} \leq \frac{y_{1}^{2}+1}{y_{1}+1} \leq y_{1} \leq x_{1}$. But this means that the pair $\left(x^{\prime}, y^{\prime}\right)$ with $x^{\prime}=y_{1}$ and $y^{\prime}=x_{2}$ is another solution of $k(x+1)(y+1)=x^{2}+y^{2}$ in even positive numbers with $x^{\prime}+y^{\prime}<x_{1}+y_{1}$, a contradiction.
Therefore we must have $n \leq 4$.
When $n=4$, a possible example of a sequence is $a_{1}=4, a_{2}=33, a_{3}=217$ and $a_{4}=1384$.

Solution 2. It is easy to check that for $n=4$ the sequence $a_{1}=4, a_{2}=33, a_{3}=217$ and $a_{4}=1384$ is possible.
Now assume there is a sequence with $n \geq 5$. Then we have in particular

$$
\begin{aligned}
a_{2}^{2}+1 & =\left(a_{1}+1\right)\left(a_{3}+1\right), \\
a_{3}^{2}+1 & =\left(a_{2}+1\right)\left(a_{4}+1\right), \\
a_{4}^{2}+1 & =\left(a_{3}+1\right)\left(a_{5}+1\right) .
\end{aligned}
$$

Also assume without loss of generality that among all such quintuples $\left(a_{1}, a_{2}, a_{3}, a_{4}, a_{5}\right)$ we have chosen one with minimal $a_{1}$.
One shows quickly the following fact:
If three positive integers $x, y, z$ fulfill $y^{2}+1=(x+1)(z+1)$ and if $y$ is even, then $x$ and $z$ are even as well and either $x<y<z$ or $z<y<x$ holds.
Indeed, the first part is obvious and from $x<y$ we conclude

$$
z+1=\frac{y^{2}+1}{x+1} \geq \frac{y^{2}+1}{y}>y
$$

and similarly in the other case.
Now, if $a_{3}$ was odd, then $\left(a_{2}+1\right)\left(a_{4}+1\right)=a_{3}^{2}+1 \equiv 2 \bmod 4$ would imply that one of $a_{2}$ or $a_{4}$ is even, this contradicts (1). Thus $a_{3}$ and hence also $a_{1}, a_{2}, a_{4}$ and $a_{5}$ are even. According to (1), one has $a_{1}<a_{2}<a_{3}<a_{4}<a_{5}$ or $a_{1}>a_{2}>a_{3}>a_{4}>a_{5}$ but due to the minimality of $a_{1}$ the first series of inequalities must hold.
Consider the identity
$\left(a_{3}+1\right)\left(a_{1}+a_{3}\right)=a_{3}^{2}-1+\left(a_{1}+1\right)\left(a_{3}+1\right)=a_{2}^{2}+a_{3}^{2}=a_{2}^{2}-1+\left(a_{2}+1\right)\left(a_{4}+1\right)=\left(a_{2}+1\right)\left(a_{2}+a_{4}\right)$.
Any common divisor of the two odd numbers $a_{2}+1$ and $a_{3}+1$ must also divide $\left(a_{2}+1\right)\left(a_{4}+\right.$ 1) $-\left(a_{3}+1\right)\left(a_{3}-1\right)=2$, so these numbers are relatively prime. Hence the last identity shows that $a_{1}+a_{3}$ must be a multiple of $a_{2}+1$, i.e. there is an integer $k$ such that

$$
\begin{equation*}
a_{1}+a_{3}=k\left(a_{2}+1\right) . \tag{2}
\end{equation*}
$$

Now set $a_{0}=k\left(a_{1}+1\right)-a_{2}$. This is an integer and we have

$$
\begin{aligned}
\left(a_{0}+1\right)\left(a_{2}+1\right) & =k\left(a_{1}+1\right)\left(a_{2}+1\right)-\left(a_{2}-1\right)\left(a_{2}+1\right) \\
& =\left(a_{1}+1\right)\left(a_{1}+a_{3}\right)-\left(a_{1}+1\right)\left(a_{3}+1\right)+2 \\
& =\left(a_{1}+1\right)\left(a_{1}-1\right)+2=a_{1}^{2}+1 .
\end{aligned}
$$

Thus $a_{0} \geq 0$. If $a_{0}>0$, then by (1) we would have $a_{0}<a_{1}<a_{2}$ and then the quintuple ( $a_{0}, a_{1}, a_{2}, a_{3}, a_{4}$ ) would contradict the minimality of $a_{1}$.
Hence $a_{0}=0$, implying $a_{2}=a_{1}^{2}$. But also $a_{2}=k\left(a_{1}+1\right)$, which finally contradicts the fact that $a_{1}+1>1$ is relatively prime to $a_{1}^{2}$ and thus cannot be a divisior of this number.
Hence $n \geq 5$ is not possible.

Comment 1. Finding the example for $n=4$ is not trivial and requires a tedious calculation, but it can be reduced to checking a few cases. The equations $\left(a_{1}+1\right)\left(a_{3}+1\right)=a_{2}^{2}+1$ and $\left(a_{2}+1\right)\left(a_{4}+1\right)=a_{3}^{2}+1$ imply, as seen in the proof, that $a_{1}$ is even and $a_{2}, a_{3}, a_{4}$ are odd. The case $a_{1}=2$ yields $a_{2}^{2} \equiv-1 \bmod 3$ which is impossible. Hence $a_{1}=4$ is the smallest possibility. In this case $a_{2}^{2} \equiv-1 \bmod 5$ and $a_{2}$ is odd, which implies $a_{2} \equiv 3$ or $a_{2} \equiv 7 \bmod 10$. Hence we have to start checking $a_{2}=7,13,17,23,27,33$ and in the last case we succeed.

Comment 2. The choice of $a_{0}=k\left(a_{1}+1\right)-a_{2}$ in the second solution appears more natural if one considers that by the previous calculations one has $a_{1}=k\left(a_{2}+1\right)-a_{3}$ and $a_{2}=k\left(a_{3}+1\right)-a_{4}$. Alternatively, one can solve the equation (2) for $a_{3}$ and use $a_{2}^{2}+1=\left(a_{1}+1\right)\left(a_{3}+1\right)$ to get $a_{2}^{2}-k\left(a_{1}+1\right) a_{2}+a_{1}^{2}-k\left(a_{1}+1\right)=0$. Now $a_{0}$ is the second solution to this quadratic equation in $a_{2}$ (Vieta jumping).

## N5 HUN (Hungary)

Let $P(x)$ be a non-constant polynomial with integer coefficients. Prove that there is no function $T$ from the set of integers into the set of integers such that the number of integers $x$ with $T^{n}(x)=x$ is equal to $P(n)$ for every $n \geq 1$, where $T^{n}$ denotes the $n$-fold application of $T$.

Solution 1. Assume there is a polynomial $P$ of degree at least 1 with the desired property for a given function $T$. Let $A(n)$ denote the set of all $x \in \mathbb{Z}$ such that $T^{n}(x)=x$ and let $B(n)$ denote the set of all $x \in \mathbb{Z}$ for which $T^{n}(x)=x$ and $T^{k}(x) \neq x$ for all $1 \leq k<n$. Both sets are finite under the assumption made. For each $x \in A(n)$ there is a smallest $k \geq 1$ such that $T^{k}(x)=x$, i.e., $x \in B(k)$. Let $d=\operatorname{gcd}(k, n)$. There are positive integers $r, s$ such that $r k-s n=d$ and hence $x=T^{r k}(x)=T^{s n+d}(x)=T^{d}\left(T^{s n}(x)\right)=T^{d}(x)$. The minimality of $k$ implies $d=k$, i.e., $k \mid n$. On the other hand one clearly has $B(k) \subset A(n)$ if $k \mid n$ and thus we have $A(n)=\bigcup_{d \mid n} B(d)$ as a disjoint union and hence

$$
|A(n)|=\sum_{d \mid n}|B(d)|
$$

Furthermore, for every $x \in B(n)$ the elements $x, T^{1}(x), T^{2}(x), \ldots, T^{n-1}(x)$ are $n$ distinct elements of $B(n)$. The fact that they are in $A(n)$ is obvious. If for some $k<n$ and some $0 \leq i<n$ we had $T^{k}\left(T^{i}(x)\right)=T^{i}(x)$, i.e. $T^{k+i}(x)=T^{i}(x)$, that would imply $x=T^{n}(x)=T^{n-i}\left(T^{i}(x)\right)=T^{n-i}\left(T^{k+i}(x)\right)=T^{k}\left(T^{n}(x)\right)=T^{k}(x)$ contradicting the minimality of $n$. Thus $T^{i}(x) \in B(n)$ and $T^{i}(x) \neq T^{j}(x)$ for $0 \leq i<j \leq n-1$.
So indeed, $T$ permutes the elements of $B(n)$ in (disjoint) cycles of length $n$ and in particular one has $n||B(n)|$.
Now let $P(x)=\sum_{i=0}^{k} a_{i} x^{i}, a_{i} \in \mathbb{Z}, k \geq 1, a_{k} \neq 0$ and suppose that $|A(n)|=P(n)$ for all $n \geq 1$. Let $p$ be any prime. Then

$$
p^{2}| | B\left(p^{2}\right)\left|=\left|A\left(p^{2}\right)\right|-|A(p)|=a_{1}\left(p^{2}-p\right)+a_{2}\left(p^{4}-p^{2}\right)+\ldots\right.
$$

Hence $p \mid a_{1}$ and since this is true for all primes we must have $a_{1}=0$.
Now consider any two different primes $p$ and $q$. Since $a_{1}=0$ we have that

$$
\left|A\left(p^{2} q\right)\right|-|A(p q)|=a_{2}\left(p^{4} q^{2}-p^{2} q^{2}\right)+a_{3}\left(p^{6} q^{3}-p^{3} q^{3}\right)+\ldots
$$

is a multiple of $p^{2} q$. But we also have

$$
p^{2} q| | B\left(p^{2} q\right)\left|=\left|A\left(p^{2} q\right)\right|-|A(p q)|-\left|B\left(p^{2}\right)\right|\right.
$$

This implies

$$
p^{2} q| | B\left(p^{2}\right)\left|=\left|A\left(p^{2}\right)\right|-|A(p)|=a_{2}\left(p^{4}-p^{2}\right)+a_{3}\left(p^{6}-p^{3}\right)+\cdots+a_{k}\left(p^{2 k}-p^{k}\right)\right.
$$

Since this is true for every prime $q$ we must have $a_{2}\left(p^{4}-p^{2}\right)+a_{3}\left(p^{6}-p^{3}\right)+\cdots+a_{k}\left(p^{2 k}-p^{k}\right)=0$ for every prime $p$. Since this expression is a polynomial in $p$ of degree $2 k$ (because $a_{k} \neq 0$ ) this is a contradiction, as such a polynomial can have at most $2 k$ zeros.

Comment. The last contradiction can also be reached via

$$
a_{k}=\lim _{p \rightarrow \infty} \frac{1}{p^{2 k}}\left(a_{2}\left(p^{4}-p^{2}\right)+a_{3}\left(p^{6}-p^{3}\right)+\cdots+a_{k}\left(p^{2 k}-p^{k}\right)\right)=0 .
$$

Solution 2. As in the first solution define $A(n)$ and $B(n)$ and assume that a polynomial $P$ with the required property exists. This again implies that $|A(n)|$ and $|B(n)|$ is finite for all positive integers $n$ and that

$$
P(n)=|A(n)|=\sum_{d \mid n}|B(d)| \quad \text { and } \quad n||B(n)| .
$$

Now, for any two distinct primes $p$ and $q$, we have

$$
P(0) \equiv P(p q) \equiv|B(1)|+|B(p)|+|B(q)|+|B(p q)| \equiv|B(1)|+|B(p)| \quad \bmod q .
$$

Thus, for any fixed $p$, the expression $P(0)-|B(1)|-|B(p)|$ is divisible by arbitrarily large primes $q$ which means that $P(0)=|B(1)|+|B(p)|=P(p)$ for any prime $p$. This implies that the polynomial $P$ is constant, a contradiction.

## N6 TUR (Turkey)

Let $k$ be a positive integer. Show that if there exists a sequence $a_{0}, a_{1}, \ldots$ of integers satisfying the condition

$$
a_{n}=\frac{a_{n-1}+n^{k}}{n} \quad \text { for all } n \geq 1,
$$

then $k-2$ is divisible by 3 .

Solution 1. Part $A$. For each positive integer $k$, there exists a polynomial $P_{k}$ of degree $k-1$ with integer coefficients, i. e., $P_{k} \in \mathbb{Z}[x]$, and an integer $q_{k}$ such that the polynomial identity

$$
\begin{equation*}
x P_{k}(x)=x^{k}+P_{k}(x-1)+q_{k} \tag{k}
\end{equation*}
$$

is satisfied. To prove this, for fixed $k$ we write

$$
P_{k}(x)=b_{k-1} x^{k-1}+\cdots+b_{1} x+b_{0}
$$

and determine the coefficients $b_{k-1}, b_{k-2}, \ldots, b_{0}$ and the number $q_{k}$ successively. Obviously, we have $b_{k-1}=1$. For $m=k-1, k-2, \ldots, 1$, comparing the coefficients of $x^{m}$ in the identity $\left(I_{k}\right)$ results in an expression of $b_{m-1}$ as an integer linear combination of $b_{k-1}, \ldots, b_{m}$, and finally $q_{k}=-P_{k}(-1)$.
Part $B$. Let $k$ be a positive integer, and let $a_{0}, a_{1}, \ldots$ be a sequence of real numbers satisfying the recursion given in the problem. This recursion can be written as

$$
a_{n}-P_{k}(n)=\frac{a_{n-1}-P_{k}(n-1)}{n}-\frac{q_{k}}{n} \quad \text { for all } n \geq 1
$$

which by induction gives

$$
a_{n}-P_{k}(n)=\frac{a_{0}-P_{k}(0)}{n!}-q_{k} \sum_{i=0}^{n-1} \frac{i!}{n!} \text { for all } n \geq 1
$$

Therefore, the numbers $a_{n}$ are integers for all $n \geq 1$ only if

$$
a_{0}=P_{k}(0) \quad \text { and } \quad q_{k}=0
$$

Part C. Multiplying the identity $\left(I_{k}\right)$ by $x^{2}+x$ and subtracting the identities $\left(I_{k+1}\right),\left(I_{k+2}\right)$ and $q_{k} x^{2}=q_{k} x^{2}$ therefrom, we obtain

$$
x T_{k}(x)=T_{k}(x-1)+2 x\left(P_{k}(x-1)+q_{k}\right)-\left(q_{k+2}+q_{k+1}+q_{k}\right),
$$

where the polynomials $T_{k} \in \mathbb{Z}[x]$ are defined by $T_{k}(x)=\left(x^{2}+x\right) P_{k}(x)-P_{k+1}(x)-P_{k+2}(x)-q_{k} x$. Thus

$$
x T_{k}(x) \equiv T_{k}(x-1)+q_{k+2}+q_{k+1}+q_{k} \bmod 2, \quad k=1,2, \ldots
$$

Comparing the degrees, we easily see that this is only possible if $T_{k}$ is the zero polynomial modulo 2 , and

$$
q_{k+2} \equiv q_{k+1}+q_{k} \bmod 2 \quad \text { for } k=1,2, \ldots
$$

Since $q_{1}=-1$ and $q_{2}=0$, these congruences finish the proof.

Solution 2. Part $A$ and $B$. Let $k$ be a positive integer, and suppose there is a sequence $a_{0}, a_{1}, \ldots$ as required. We prove: There exists a polynomial $P \in \mathbb{Z}[x]$, i. e., with integer coefficients, such that $a_{n}=P(n), n=0,1, \ldots$, and $\quad x P(x)=x^{k}+P(x-1)$.
To prove this, we write $P(x)=b_{k-1} x^{k-1}+\cdots+b_{1} x+b_{0} \quad$ and determine the coefficients $b_{k-1}, b_{k-2}, \ldots, b_{0}$ successively such that

$$
x P(x)-x^{k}-P(x-1)=q,
$$

where $q=q_{k}$ is an integer. Comparing the coefficients of $x^{m}$ results in an expression of $b_{m-1}$ as an integer linear combination of $b_{k-1}, \ldots, b_{m}$.
Defining $c_{n}=a_{n}-P(n)$, we get

$$
\begin{aligned}
P(n)+c_{n} & =\frac{P(n-1)+c_{n-1}+n^{k}}{n}, \quad \text { i. e., } \\
q+n c_{n} & =c_{n-1},
\end{aligned}
$$

hence

$$
c_{n}=\frac{c_{0}}{n!}-q \cdot \frac{0!+1!+\cdots+(n-1)!}{n!}
$$

We conclude $\lim _{n \rightarrow \infty} c_{n}=0$, which, using $c_{n} \in \mathbb{Z}$, implies $c_{n}=0$ for sufficiently large $n$. Therefore, we get $q=0$ and $c_{n}=0, n=0,1, \ldots$.
Part C. Suppose that $q=q_{k}=0$, i. e. $x P(x)=x^{k}+P(x-1)$. To consider this identity for arguments $x \in \mathbb{F}_{4}$, we write $\mathbb{F}_{4}=\{0,1, \alpha, \alpha+1\}$. Then we get

$$
\begin{aligned}
\alpha P_{k}(\alpha) & =\alpha^{k}+P_{k}(\alpha+1) \quad \text { and } \\
(\alpha+1) P_{k}(\alpha+1) & =(\alpha+1)^{k}+P_{k}(\alpha),
\end{aligned}
$$

hence

$$
\begin{aligned}
P_{k}(\alpha) & =1 \cdot P_{k}(\alpha)=(\alpha+1) \alpha P_{k}(\alpha) \\
& =(\alpha+1) P_{k}(\alpha+1)+(\alpha+1) \alpha^{k} \\
& =P_{k}(\alpha)+(\alpha+1)^{k}+(\alpha+1) \alpha^{k} .
\end{aligned}
$$

Now, $(\alpha+1)^{k-1}=\alpha^{k}$ implies $k \equiv 2 \bmod 3$.

Comment 1. For $k=2$, the sequence given by $a_{n}=n+1, n=0,1, \ldots$, satisfies the conditions of the problem.

Comment 2. The first few polynomials $P_{k}$ and integers $q_{k}$ are

$$
\begin{aligned}
& P_{1}(x)=1, \quad q_{1}=-1, \\
& P_{2}(x)=x+1, \quad q_{2}=0, \\
& P_{3}(x)=x^{2}+x-1, \quad q_{3}=1, \\
& P_{4}(x)=x^{3}+x^{2}-2 x-1, \quad q_{4}=-1, \\
& P_{5}(x)=x^{4}+x^{3}-3 x^{2}+5, \quad q_{5}=-2, \\
& P_{6}(x)=x^{5}+x^{4}-4 x^{3}+2 x^{2}+10 x-5, \quad q_{6}=9, \\
& q_{7}=-9, \quad q_{8}=-50, \quad q_{9}=267, \quad q_{10}=-413, \quad q_{11}=-2180 .
\end{aligned}
$$

A lookup in the On-Line Encyclopedia of Integer Sequences (A000587) reveals that the sequence $q_{1},-q_{2}, q_{3},-q_{4}, q_{5}, \ldots$ is known as Uppuluri-Carpenter numbers. The result that $q_{k}=0$ implies $k \equiv 2 \bmod 3$ is contained in
Murty, Summer: On the $p$-adic series $\sum_{n=0}^{\infty} n^{k} \cdot n$ !. CRM Proc. and Lecture Notes 36, 2004. As shown by Alexander (Non-Vanishing of Uppuluri-Carpenter Numbers, Preprint 2006), Uppuluri-Carpenter numbers are zero at most twice.

Comment 3. The numbers $q_{k}$ can be written in terms of the Stirling numbers of the second kind. To show this, we fix the notation such that

$$
\begin{align*}
x^{k}= & S_{k-1, k-1} x(x-1) \cdots(x-k+1) \\
& +S_{k-1, k-2} x(x-1) \cdots(x-k+2)  \tag{*}\\
& +\cdots+S_{k-1,0} x,
\end{align*}
$$

e.g., $S_{2,2}=1, S_{2,1}=3, S_{2,0}=1$, and we define

$$
\Omega_{k}=S_{k-1, k-1}-S_{k-1, k-2}+-\cdots
$$

Replacing $x$ by $-x$ in (*) results in

$$
\begin{aligned}
x^{k}= & S_{k-1, k-1} x(x+1) \cdots(x+k-1) \\
& -S_{k-1, k-2} x(x+1) \cdots(x+k-2) \\
& +-\cdots \pm S_{k-1,0} x .
\end{aligned}
$$

Defining

$$
\begin{aligned}
P(x)= & S_{k-1, k-1}(x+1) \cdots(x+k-1) \\
& +\left(S_{k-1, k-1}-S_{k-1, k-2}\right)(x+1) \cdots(x+k-2) \\
& +\left(S_{k-1, k-1}-S_{k-1, k-2}+S_{k-1, k-3}\right)(x+1) \cdots(x+k-3) \\
& +\cdots+\Omega_{k},
\end{aligned}
$$

we obtain

$$
\begin{aligned}
x P(x)-P(x-1)= & S_{k-1, k-1} x(x+1) \cdots(x+k-1) \\
& -S_{k-1, k-2} x(x+1) \cdots(x+k-2) \\
& +-\cdots \pm S_{k-1,0} x-\Omega_{k} \\
= & x^{k}-\Omega_{k},
\end{aligned}
$$

hence $q_{k}=-\Omega_{k}$.

## N7 MNG (Mongolia)

Let $a$ and $b$ be distinct integers greater than 1 . Prove that there exists a positive integer $n$ such that $\left(a^{n}-1\right)\left(b^{n}-1\right)$ is not a perfect square.

Solution 1. At first we notice that

$$
\begin{align*}
(1-\alpha)^{\frac{1}{2}}(1-\beta)^{\frac{1}{2}} & =\left(1-\frac{1}{2} \cdot \alpha-\frac{1}{8} \cdot \alpha^{2}-\cdots\right)\left(1-\frac{1}{2} \cdot \beta-\frac{1}{8} \cdot \beta^{2}-\cdots\right) \\
& =\sum_{k, \ell \geq 0} c_{k, \ell} \cdot \alpha^{k} \beta^{\ell} \quad \text { for all } \alpha, \beta \in(0,1) \tag{1}
\end{align*}
$$

where $c_{0,0}=1$ and $c_{k, \ell}$ are certain coefficients.
For an indirect proof, we suppose that $x_{n}=\sqrt{\left(a^{n}-1\right)\left(b^{n}-1\right)} \in \mathbb{Z}$ for all positive integers $n$. Replacing $a$ by $a^{2}$ and $b$ by $b^{2}$ if necessary, we may assume that $a$ and $b$ are perfect squares, hence $\sqrt{a b}$ is an integer.
At first we shall assume that $a^{\mu} \neq b^{\nu}$ for all positive integers $\mu, \nu$. We have

$$
\begin{equation*}
x_{n}=(\sqrt{a b})^{n}\left(1-\frac{1}{a^{n}}\right)^{\frac{1}{2}}\left(1-\frac{1}{b^{n}}\right)^{\frac{1}{2}}=\sum_{k, \ell \geq 0} c_{k, \ell}\left(\frac{\sqrt{a b}}{a^{k} b^{\ell}}\right)^{n} . \tag{2}
\end{equation*}
$$

Choosing $k_{0}$ and $\ell_{0}$ such that $a^{k_{0}}>\sqrt{a b}, b^{\ell_{0}}>\sqrt{a b}$, we define the polynomial

$$
P(x)=\prod_{k=0, \ell=0}^{k_{0}-1, \ell_{0}-1}\left(a^{k} b^{\ell} x-\sqrt{a b}\right)=: \sum_{i=0}^{k_{0} \cdot \ell_{0}} d_{i} x^{i}
$$

with integer coefficients $d_{i}$. By our assumption, the zeros

$$
\frac{\sqrt{a b}}{a^{k} b^{\ell}}, \quad k=0, \ldots, k_{0}-1, \quad \ell=0, \ldots, \ell_{0}-1,
$$

of $P$ are pairwise distinct.
Furthermore, we consider the integer sequence

$$
\begin{equation*}
y_{n}=\sum_{i=0}^{k_{0} \cdot \ell_{0}} d_{i} x_{n+i}, \quad n=1,2, \ldots \tag{3}
\end{equation*}
$$

By the theory of linear recursions, we obtain

$$
\begin{equation*}
y_{n}=\sum_{\substack{k, \ell \geq 0 \\ k \geq k_{0} \text { or } \ell \geq \ell_{0}}} e_{k, \ell}\left(\frac{\sqrt{a b}}{a^{k} b^{\ell}}\right)^{n}, \quad n=1,2, \ldots, \tag{4}
\end{equation*}
$$

with real numbers $e_{k, \ell}$. We have

$$
\left|y_{n}\right| \leq \sum_{\substack{k, \ell \geq 0 \\ k \geq k_{0} \text { or } \ell \geq \ell_{0}}}\left|e_{k, \ell}\right|\left(\frac{\sqrt{a b}}{a^{k} b^{\ell}}\right)^{n}=: M_{n} .
$$

Because the series in (4) is obtained by a finite linear combination of the absolutely convergent series (1), we conclude that in particular $M_{1}<\infty$. Since

$$
\frac{\sqrt{a b}}{a^{k} b^{\ell}} \leq \lambda:=\max \left\{\frac{\sqrt{a b}}{a^{k_{0}}}, \frac{\sqrt{a b}}{b^{\ell_{0}}}\right\} \quad \text { for all } k, \ell \geq 0 \text { such that } k \geq k_{0} \text { or } \ell \geq \ell_{0}
$$

we get the estimates $M_{n+1} \leq \lambda M_{n}, n=1,2, \ldots$ Our choice of $k_{0}$ and $\ell_{0}$ ensures $\lambda<1$, which implies $M_{n} \rightarrow 0$ and consequently $y_{n} \rightarrow 0$ as $n \rightarrow \infty$. It follows that $y_{n}=0$ for all sufficiently large $n$.
So, equation (3) reduces to $\sum_{i=0}^{k_{0} \cdot \ell_{0}} d_{i} x_{n+i}=0$.
Using the theory of linear recursions again, for sufficiently large $n$ we have

$$
x_{n}=\sum_{k=0, \ell=0}^{k_{0}-1, \ell_{0}-1} f_{k, \ell}\left(\frac{\sqrt{a b}}{a^{k} b^{\ell}}\right)^{n}
$$

for certain real numbers $f_{k, \ell}$.
Comparing with (2), we see that $f_{k, \ell}=c_{k, \ell}$ for all $k, \ell \geq 0$ with $k<k_{0}, \ell<\ell_{0}$, and $c_{k, \ell}=0$ if $k \geq k_{0}$ or $\ell \geq \ell_{0}$, since we assumed that $a^{\mu} \neq b^{\nu}$ for all positive integers $\mu, \nu$.
In view of (1), this means

$$
\begin{equation*}
(1-\alpha)^{\frac{1}{2}}(1-\beta)^{\frac{1}{2}}=\sum_{k=0, \ell=0}^{k_{0}-1, \ell_{0}-1} c_{k, \ell} \cdot \alpha^{k} \beta^{\ell} \tag{5}
\end{equation*}
$$

for all real numbers $\alpha, \beta \in(0,1)$. We choose $k^{*}<k_{0}$ maximal such that there is some $i$ with $c_{k^{*}, i} \neq 0$. Squaring (5) and comparing coefficients of $\alpha^{2 k^{*}} \beta^{2 i^{*}}$, where $i^{*}$ is maximal with $c_{k^{*}, i^{*}} \neq 0$, we see that $k^{*}=0$. This means that the right hand side of (5) is independent of $\alpha$, which is clearly impossible.
We are left with the case that $a^{\mu}=b^{\nu}$ for some positive integers $\mu$ and $\nu$. We may assume that $\mu$ and $\nu$ are relatively prime. Then there is some positive integer $c$ such that $a=c^{\nu}$ and $b=c^{\mu}$. Now starting with the expansion (2), i. e.,

$$
x_{n}=\sum_{j \geq 0} g_{j}\left(\frac{\sqrt{c^{\mu+\nu}}}{c^{j}}\right)^{n}
$$

for certain coefficients $g_{j}$, and repeating the arguments above, we see that $g_{j}=0$ for sufficiently large $j$, say $j>j_{0}$. But this means that

$$
\left(1-x^{\mu}\right)^{\frac{1}{2}}\left(1-x^{\nu}\right)^{\frac{1}{2}}=\sum_{j=0}^{j_{0}} g_{j} x^{j}
$$

for all real numbers $x \in(0,1)$. Squaring, we see that

$$
\left(1-x^{\mu}\right)\left(1-x^{\nu}\right)
$$

is the square of a polynomial in $x$. In particular, all its zeros are of order at least 2 , which implies $\mu=\nu$ by looking at roots of unity. So we obtain $\mu=\nu=1$, i. e., $a=b$, a contradiction.

Solution 2. We set $a^{2}=A, b^{2}=B$, and $z_{n}=\sqrt{\left(A^{n}-1\right)\left(B^{n}-1\right)}$. Let us assume that $z_{n}$ is an integer for $n=1,2, \ldots$. Without loss of generality, we may suppose that $b<a$. We determine an integer $k \geq 2$ such that $b^{k-1} \leq a<b^{k}$, and define a sequence $\gamma_{1}, \gamma_{2}, \ldots$ of rational numbers such that

$$
2 \gamma_{1}=1 \quad \text { and } \quad 2 \gamma_{n+1}=\sum_{i=1}^{n} \gamma_{i} \gamma_{n-i} \text { for } n=1,2, \ldots
$$

It could easily be shown that $\gamma_{n}=\frac{1 \cdot 1 \cdot 3 \cdot . .(2 n-3)}{2 \cdot 4 \cdot 6 \ldots 2 n}$, for instance by reading Vandermondes convolution as an equation between polynomials, but we shall have no use for this fact.
Using Landaus $O$-Notation in the usual way, we have

$$
\begin{aligned}
& \left\{(a b)^{n}-\gamma_{1}\left(\frac{a}{b}\right)^{n}-\gamma_{2}\left(\frac{a}{b^{3}}\right)^{n}-\cdots-\gamma_{k}\left(\frac{a}{b^{2 k-1}}\right)^{n}+O\left(\frac{b}{a}\right)^{n}\right\}^{2} \\
& =A^{n} B^{n}-2 \gamma_{1} A^{n}-\sum_{i=2}^{k}\left(2 \gamma_{i}-\sum_{j=1}^{i-1} \gamma_{j} \gamma_{i-j}\right)\left(\frac{A}{B^{i-1}}\right)^{n}+O\left(\frac{A}{B^{k}}\right)^{n}+O\left(B^{n}\right) \\
& =A^{n} B^{n}-A^{n}+O\left(B^{n}\right)
\end{aligned}
$$

whence

$$
z_{n}=(a b)^{n}-\gamma_{1}\left(\frac{a}{b}\right)^{n}-\gamma_{2}\left(\frac{a}{b^{3}}\right)^{n}-\cdots-\gamma_{k}\left(\frac{a}{b^{2 k-1}}\right)^{n}+O\left(\frac{b}{a}\right)^{n} .
$$

Now choose rational numbers $r_{1}, r_{2}, \ldots, r_{k+1}$ such that

$$
(x-a b) \cdot\left(x-\frac{a}{b}\right) \ldots\left(x-\frac{a}{b^{2 k-1}}\right)=x^{k+1}-r_{1} x^{k}+-\cdots \pm r_{k+1},
$$

and then a natural number $M$ for which $M r_{1}, M r_{2}, \ldots M r_{k+1}$ are integers. For known reasons,

$$
M\left(z_{n+k+1}-r_{1} z_{n+k}+-\cdots \pm r_{k+1} z_{n}\right)=O\left(\frac{b}{a}\right)^{n}
$$

for all $n \in \mathbb{N}$ and thus there is a natural number $N$ which is so large, that

$$
z_{n+k+1}=r_{1} z_{n+k}-r_{2} z_{n+k-1}+-\cdots \mp r_{k+1} z_{n}
$$

holds for all $n \geqslant N$. Now the theory of linear recursions reveals that there are some rational numbers $\delta_{0}, \delta_{1}, \delta_{2}, \ldots, \delta_{k}$ such that

$$
z_{n}=\delta_{0}(a b)^{n}-\delta_{1}\left(\frac{a}{b}\right)^{n}-\delta_{2}\left(\frac{a}{b^{3}}\right)^{n}-\cdots-\delta_{k}\left(\frac{a}{b^{2 k-1}}\right)^{n}
$$

for sufficiently large $n$, where $\delta_{0}>0$ as $z_{n}>0$. As before, one obtains

$$
\begin{aligned}
& A^{n} B^{n}-A^{n}-B^{n}+1=z_{n}^{2} \\
& =\left\{\delta_{0}(a b)^{n}-\delta_{1}\left(\frac{a}{b}\right)^{n}-\delta_{2}\left(\frac{a}{b^{3}}\right)^{n}-\cdots-\delta_{k}\left(\frac{a}{b^{2 k-1}}\right)^{n}\right\}^{2} \\
& =\delta_{0}^{2} A^{n} B^{n}-2 \delta_{0} \delta_{1} A^{n}-\sum_{i=2}^{i=k}\left(2 \delta_{0} \delta_{i}-\sum_{j=1}^{j=i-1} \delta_{j} \delta_{i-j}\right)\left(\frac{A}{B^{i-1}}\right)^{n}+O\left(\frac{A}{B^{k}}\right)^{n} .
\end{aligned}
$$

Easy asymptotic calculations yield $\delta_{0}=1, \delta_{1}=\frac{1}{2}, \delta_{i}=\frac{1}{2} \sum_{j=1}^{j=i-1} \delta_{j} \delta_{i-j}$ for $i=2,3, \ldots, k-2$, and then $a=b^{k-1}$. It follows that $k>2$ and there is some $P \in \mathbb{Q}[X]$ for which $(X-1)\left(X^{k-1}-1\right)=$ $P(X)^{2}$. But this cannot occur, for instance as $X^{k-1}-1$ has no double zeros. Thus our
assumption that $z_{n}$ was an integer for $n=1,2, \ldots$ turned out to be wrong, which solves the problem.

Original formulation of the problem. $a, b$ are positive integers such that $a \cdot b$ is not a square of an integer. Prove that there exists a (infinitely many) positive integer $n$ such that ( $\left.a^{n}-1\right)\left(b^{n}-1\right)$ is not a square of an integer.

Solution. Lemma. Let $c$ be a positive integer, which is not a perfect square. Then there exists an odd prime $p$ such that $c$ is not a quadratic residue modulo $p$.
Proof. Denoting the square-free part of $c$ by $c^{\prime}$, we have the equality $\left(\frac{c^{\prime}}{p}\right)=\left(\frac{c}{p}\right)$ of the corresponding Legendre symbols. Suppose that $c^{\prime}=q_{1} \cdots q_{m}$, where $q_{1}<\cdots<q_{m}$ are primes. Then we have

$$
\left(\frac{c^{\prime}}{p}\right)=\left(\frac{q_{1}}{p}\right) \cdots\left(\frac{q_{m}}{p}\right)
$$

Case 1. Let $q_{1}$ be odd. We choose a quadratic nonresidue $r_{1}$ modulo $q_{1}$ and quadratic residues $r_{i}$ modulo $q_{i}$ for $i=2, \ldots, m$. By the Chinese remainder theorem and the Dirichlet theorem, there exists a (infinitely many) prime $p$ such that

$$
\begin{aligned}
& p \equiv r_{1} \bmod q_{1} \\
& p \equiv r_{2} \bmod q_{2} \\
& \vdots \vdots \\
& p \equiv r_{m} \bmod q_{m}, \\
& p \equiv 1 \bmod 4
\end{aligned}
$$

By our choice of the residues, we obtain

$$
\left(\frac{p}{q_{i}}\right)=\left(\frac{r_{i}}{q_{i}}\right)= \begin{cases}-1, & i=1 \\ 1, & i=2, \ldots, m\end{cases}
$$

The congruence $p \equiv 1 \bmod 4$ implies that $\left(\frac{q_{i}}{p}\right)=\left(\frac{p}{q_{i}}\right), i=1, \ldots, m$, by the law of quadratic reciprocity. Thus

$$
\left(\frac{c^{\prime}}{p}\right)=\left(\frac{q_{1}}{p}\right) \cdots\left(\frac{q_{m}}{p}\right)=-1 .
$$

Case 2. Suppose $q_{1}=2$. We choose quadratic residues $r_{i}$ modulo $q_{i}$ for $i=2, \ldots, m$. Again, by the Chinese remainder theorem and the Dirichlet theorem, there exists a prime $p$ such that

$$
\begin{aligned}
& p \equiv r_{2} \bmod q_{2} \\
& \vdots \quad \vdots \\
& p \equiv r_{m} \bmod q_{m} \\
& p \equiv 5 \bmod 8
\end{aligned}
$$

By the choice of the residues, we obtain $\left(\frac{p}{q_{i}}\right)=\left(\frac{r_{i}}{q_{i}}\right)=1$ for $i=2, \ldots, m$. Since $p \equiv 1 \bmod 4$ we have $\left(\frac{q_{i}}{p}\right)=\left(\frac{p}{q_{i}}\right), i=2, \ldots, m$, by the law of quadratic reciprocity. The congruence $p \equiv 5 \bmod 8$
implies that $\left(\frac{2}{p}\right)=-1$. Thus

$$
\left(\frac{c^{\prime}}{p}\right)=\left(\frac{2}{p}\right)\left(\frac{q_{2}}{p}\right) \cdots\left(\frac{q_{m}}{p}\right)=-1
$$

and the lemma is proved.
Applying the lemma for $c=a \cdot b$, we find an odd prime $p$ such that

$$
\left(\frac{a b}{p}\right)=\left(\frac{a}{p}\right) \cdot\left(\frac{b}{p}\right)=-1
$$

This implies either

$$
a^{\frac{p-1}{2}} \equiv 1 \bmod p, \quad b^{\frac{p-1}{2}} \equiv-1 \bmod p, \quad \text { or } \quad a^{\frac{p-1}{2}} \equiv-1 \bmod p, \quad b^{\frac{p-1}{2}} \equiv 1 \bmod p
$$

Without loss of generality, suppose that $a^{\frac{p-1}{2}} \equiv 1 \bmod p$ and $b^{\frac{p-1}{2}} \equiv-1 \bmod p$. The second congruence implies that $b^{\frac{p-1}{2}}-1$ is not divisible by $p$. Hence, if the exponent $\nu_{p}\left(a^{\frac{p-1}{2}}-1\right)$ of $p$ in the prime decomposition of $\left(a^{\frac{p-1}{2}}-1\right)$ is odd, then $\left(a^{\frac{p-1}{2}}-1\right)\left(b^{\frac{p-1}{2}}-1\right)$ is not a perfect square. If $\nu_{p}\left(a^{\frac{p-1}{2}}-1\right)$ is even, then $\nu_{p}\left(a^{\frac{p-1}{2} p}-1\right)$ is odd by the well-known power lifting property

$$
\nu_{p}\left(a^{\frac{p-1}{2} p}-1\right)=\nu_{p}\left(a^{\frac{p-1}{2}}-1\right)+1 .
$$

In this case, $\left(a^{\frac{p-1}{2} p}-1\right)\left(b^{\frac{p-1}{2} p}-1\right)$ is not a perfect square.

Comment 1. In 1998, the following problem appeared in Crux Mathematicorum:
Problem 2344. Find all positive integers $N$ that are quadratic residues modulo all primes greater than $N$.
The published solution (Crux Mathematicorum, 25(1999)4) is the same as the proof of the lemma given above, see also http://www.mathlinks.ro/viewtopic.php?t=150495.

Comment 2. There is also an elementary proof of the lemma. We cite Theorem 3 of Chapter 5 and its proof from the book
Ireland, Rosen: A Classical Introduction to Modern Number Theory, Springer 1982.
Theorem. Let $a$ be a nonsquare integer. Then there are infinitely many primes $p$ for which $a$ is a quadratic nonresidue.
Proof. It is easily seen that we may assume that $a$ is square-free. Let $a=2^{e} q_{1} q_{2} \cdots q_{n}$, where $q_{i}$ are distinct odd primes and $e=0$ or 1 . The case $a=2$ has to be dealt with separately. We shall assume to begin with that $n \geq 1$, i. e., that $a$ is divisible by an odd prime.

Let $\ell_{1}, \ell_{2}, \ldots, \ell_{k}$ be a finite set of odd primes not including any $q_{i}$. Let $s$ be any quadratic nonresidue $\bmod q_{n}$, and find a simultaneous solution to the congruences

$$
\begin{aligned}
& x \equiv 1 \bmod \ell_{i}, \quad i=1, \ldots, k, \\
& x \equiv 1 \bmod 8, \\
& x \equiv 1 \bmod q_{i}, \quad i=1, \ldots, n-1, \\
& x \equiv s \bmod q_{n} .
\end{aligned}
$$

Call the solution $b . b$ is odd. Suppose that $b=p_{1} p_{2} \cdots p_{m}$ is its prime decomposition. Since
$b \equiv 1 \bmod 8$ we have $\left(\frac{2}{b}\right)=1$ and $\left(\frac{q_{i}}{b}\right)=\left(\frac{b}{q_{i}}\right)$ by a result on JACOBI symbols. Thus

$$
\left(\frac{a}{b}\right)=\left(\frac{2}{b}\right)^{e}\left(\frac{q_{1}}{b}\right) \cdots\left(\frac{q_{n-1}}{b}\right)\left(\frac{q_{n}}{b}\right)=\left(\frac{b}{q_{1}}\right) \cdots\left(\frac{b}{q_{n-1}}\right)\left(\frac{b}{q_{n}}\right)=\left(\frac{1}{q_{1}}\right) \cdots\left(\frac{1}{q_{n-1}}\right)\left(\frac{s}{q_{n}}\right)=-1 .
$$

On the other hand, by the definition of $\left(\frac{a}{b}\right)$, we have $\left(\frac{a}{b}\right)=\left(\frac{a}{p_{1}}\right)\left(\frac{a}{p_{2}}\right) \cdots\left(\frac{a}{p_{m}}\right)$. It follows that $\left(\frac{a}{p_{i}}\right)=-1$ for some $i$.
Notice that $\ell_{j}$ does not divide $b$. Thus $p_{i} \notin\left\{\ell_{1}, \ell_{2}, \ldots, \ell_{k}\right\}$.
To summarize, if $a$ is a nonsquare, divisible by an odd prime, we have found a prime $p$, outside of a given finite set of primes $\left\{2, \ell_{1}, \ell_{2}, \ldots, \ell_{k}\right\}$, such that $\left(\frac{a}{p}\right)=-1$. This proves the theorem in this case.
It remains to consider the case $a=2$. Let $\ell_{1}, \ell_{2}, \ldots, \ell_{k}$ be a finite set of primes, excluding 3 , for which $\left(\frac{2}{\ell_{i}}\right)=-1$. Let $b=8 \ell_{1} \ell_{2} \cdots \ell_{k}+3$. $b$ is not divisible by 3 or any $\ell_{i}$. Since $b \equiv 3 \bmod 8$ we have $\left(\frac{2}{b}\right)=(-1)^{\frac{b^{2}-1}{8}}=-1$. Suppose that $b=p_{1} p_{2} \cdots p_{m}$ is the prime decomposition of $b$. Then, as before, we see that $\left(\frac{2}{p_{i}}\right)=-1$ for some $i$. $p_{i} \notin\left\{3, \ell_{1}, \ell_{2}, \ldots, \ell_{k}\right\}$. This proves the theorem for $a=2$.
This proof has also been posted to mathlinks, see http://www.mathlinks.ro/viewtopic. php? $t=150495$ again.

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51st IMO Shortlisted Problems with Solutions
$\square$
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## Shortlisted Problems with Solutions

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## Note of Confidentiality

## The Shortlisted Problems should be kept strictly confidential until IMO 2011.

## Contributing Countries

The Organizing Committee and the Problem Selection Committee of IMO 2010 thank the following 42 countries for contributing 158 problem proposals.

Armenia, Australia, Austria, Bulgaria, Canada, Columbia, Croatia, Cyprus, Estonia, Finland, France, Georgia, Germany, Greece, Hong Kong, Hungary, India, Indonesia, Iran, Ireland, Japan, Korea (North), Korea (South), Luxembourg, Mongolia, Netherlands, Pakistan, Panama, Poland, Romania, Russia, Saudi Arabia, Serbia, Slovakia, Slovenia, Switzerland, Thailand, Turkey, Ukraine, United Kingdom, United States of America, Uzbekistan

## Problem Selection Committee

Yerzhan Baissalov
Ilya Bogdanov
Géza Kós
Nairi Sedrakyan
Damir Yeliussizov
Kuat Yessenov

## Algebra

A1. Determine all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ such that the equality

$$
\begin{equation*}
f([x] y)=f(x)[f(y)] . \tag{1}
\end{equation*}
$$

holds for all $x, y \in \mathbb{R}$. Here, by $[x]$ we denote the greatest integer not exceeding $x$.
(France)
Answer. $f(x)=$ const $=C$, where $C=0$ or $1 \leq C<2$.
Solution 1. First, setting $x=0$ in (1) we get

$$
\begin{equation*}
f(0)=f(0)[f(y)] \tag{2}
\end{equation*}
$$

for all $y \in \mathbb{R}$. Now, two cases are possible.
Case 1. Assume that $f(0) \neq 0$. Then from (2) we conclude that $[f(y)]=1$ for all $y \in \mathbb{R}$. Therefore, equation (1) becomes $f([x] y)=f(x)$, and substituting $y=0$ we have $f(x)=f(0)=C \neq 0$. Finally, from $[f(y)]=1=[C]$ we obtain that $1 \leq C<2$.

Case 2. Now we have $f(0)=0$. Here we consider two subcases.
Subcase 2a. Suppose that there exists $0<\alpha<1$ such that $f(\alpha) \neq 0$. Then setting $x=\alpha$ in (1) we obtain $0=f(0)=f(\alpha)[f(y)]$ for all $y \in \mathbb{R}$. Hence, $[f(y)]=0$ for all $y \in \mathbb{R}$. Finally, substituting $x=1$ in (1) provides $f(y)=0$ for all $y \in \mathbb{R}$, thus contradicting the condition $f(\alpha) \neq 0$.

Subcase 2b. Conversely, we have $f(\alpha)=0$ for all $0 \leq \alpha<1$. Consider any real $z$; there exists an integer $N$ such that $\alpha=\frac{z}{N} \in[0,1)$ (one may set $N=[z]+1$ if $z \geq 0$ and $N=[z]-1$ otherwise). Now, from (1) we get $f(z)=f([N] \alpha)=f(N)[f(\alpha)]=0$ for all $z \in \mathbb{R}$.

Finally, a straightforward check shows that all the obtained functions satisfy (1).
Solution 2. Assume that $[f(y)]=0$ for some $y$; then the substitution $x=1$ provides $f(y)=f(1)[f(y)]=0$. Hence, if $[f(y)]=0$ for all $y$, then $f(y)=0$ for all $y$. This function obviously satisfies the problem conditions.

So we are left to consider the case when $[f(a)] \neq 0$ for some $a$. Then we have

$$
\begin{equation*}
f([x] a)=f(x)[f(a)], \quad \text { or } \quad f(x)=\frac{f([x] a)}{[f(a)]} . \tag{3}
\end{equation*}
$$

This means that $f\left(x_{1}\right)=f\left(x_{2}\right)$ whenever $\left[x_{1}\right]=\left[x_{2}\right]$, hence $f(x)=f([x])$, and we may assume that $a$ is an integer.

Now we have

$$
f(a)=f\left(2 a \cdot \frac{1}{2}\right)=f(2 a)\left[f\left(\frac{1}{2}\right)\right]=f(2 a)[f(0)] ;
$$

this implies $[f(0)] \neq 0$, so we may even assume that $a=0$. Therefore equation (3) provides

$$
f(x)=\frac{f(0)}{[f(0)]}=C \neq 0
$$

for each $x$. Now, condition (1) becomes equivalent to the equation $C=C[C]$ which holds exactly when $[C]=1$.

A2. Let the real numbers $a, b, c, d$ satisfy the relations $a+b+c+d=6$ and $a^{2}+b^{2}+c^{2}+d^{2}=12$. Prove that

$$
36 \leq 4\left(a^{3}+b^{3}+c^{3}+d^{3}\right)-\left(a^{4}+b^{4}+c^{4}+d^{4}\right) \leq 48
$$

(Ukraine)
Solution 1. Observe that

$$
\begin{gathered}
4\left(a^{3}+b^{3}+c^{3}+d^{3}\right)-\left(a^{4}+b^{4}+c^{4}+d^{4}\right)=-\left((a-1)^{4}+(b-1)^{4}+(c-1)^{4}+(d-1)^{4}\right) \\
+6\left(a^{2}+b^{2}+c^{2}+d^{2}\right)-4(a+b+c+d)+4 \\
=-\left((a-1)^{4}+(b-1)^{4}+(c-1)^{4}+(d-1)^{4}\right)+52
\end{gathered}
$$

Now, introducing $x=a-1, y=b-1, z=c-1, t=d-1$, we need to prove the inequalities

$$
16 \geq x^{4}+y^{4}+z^{4}+t^{4} \geq 4,
$$

under the constraint

$$
\begin{equation*}
x^{2}+y^{2}+z^{2}+t^{2}=\left(a^{2}+b^{2}+c^{2}+d^{2}\right)-2(a+b+c+d)+4=4 \tag{1}
\end{equation*}
$$

(we will not use the value of $x+y+z+t$ though it can be found).
Now the rightmost inequality in (1) follows from the power mean inequality:

$$
x^{4}+y^{4}+z^{4}+t^{4} \geq \frac{\left(x^{2}+y^{2}+z^{2}+t^{2}\right)^{2}}{4}=4 .
$$

For the other one, expanding the brackets we note that

$$
\left(x^{2}+y^{2}+z^{2}+t^{2}\right)^{2}=\left(x^{4}+y^{4}+z^{4}+t^{4}\right)+q,
$$

where $q$ is a nonnegative number, so

$$
x^{4}+y^{4}+z^{4}+t^{4} \leq\left(x^{2}+y^{2}+z^{2}+t^{2}\right)^{2}=16
$$

and we are done.
Comment 1. The estimates are sharp; the lower and upper bounds are attained at ( $3,1,1,1$ ) and $(0,2,2,2)$, respectively.

Comment 2. After the change of variables, one can finish the solution in several different ways. The latter estimate, for instance, it can be performed by moving the variables - since we need only the second of the two shifted conditions.

Solution 2. First, we claim that $0 \leq a, b, c, d \leq 3$. Actually, we have

$$
a+b+c=6-d, \quad a^{2}+b^{2}+c^{2}=12-d^{2}
$$

hence the power mean inequality

$$
a^{2}+b^{2}+c^{2} \geq \frac{(a+b+c)^{2}}{3}
$$

rewrites as

$$
12-d^{2} \geq \frac{(6-d)^{2}}{3} \quad \Longleftrightarrow \quad 2 d(d-3) \leq 0
$$

which implies the desired inequalities for $d$; since the conditions are symmetric, we also have the same estimate for the other variables.

Now, to prove the rightmost inequality, we use the obvious inequality $x^{2}(x-2)^{2} \geq 0$ for each real $x$; this inequality rewrites as $4 x^{3}-x^{4} \leq 4 x^{2}$. It follows that

$$
\left(4 a^{3}-a^{4}\right)+\left(4 b^{3}-b^{4}\right)+\left(4 c^{3}-c^{4}\right)+\left(4 d^{3}-d^{4}\right) \leq 4\left(a^{2}+b^{2}+c^{2}+d^{2}\right)=48
$$

as desired.
Now we prove the leftmost inequality in an analogous way. For each $x \in[0,3]$, we have $(x+1)(x-1)^{2}(x-3) \leq 0$ which is equivalent to $4 x^{3}-x^{4} \geq 2 x^{2}+4 x-3$. This implies that
$\left(4 a^{3}-a^{4}\right)+\left(4 b^{3}-b^{4}\right)+\left(4 c^{3}-c^{4}\right)+\left(4 d^{3}-d^{4}\right) \geq 2\left(a^{2}+b^{2}+c^{2}+d^{2}\right)+4(a+b+c+d)-12=36$, as desired.

Comment. It is easy to guess the extremal points $(0,2,2,2)$ and $(3,1,1,1)$ for this inequality. This provides a method of finding the polynomials used in Solution 2. Namely, these polynomials should have the form $x^{4}-4 x^{3}+a x^{2}+b x+c$; moreover, the former polynomial should have roots at 2 (with an even multiplicity) and 0 , while the latter should have roots at 1 (with an even multiplicity) and 3 . These conditions determine the polynomials uniquely.

Solution 3. First, expanding $48=4\left(a^{2}+b^{2}+c^{2}+d^{2}\right)$ and applying the AM-GM inequality, we have

$$
\begin{aligned}
a^{4}+b^{4}+c^{4}+d^{4}+48 & =\left(a^{4}+4 a^{2}\right)+\left(b^{4}+4 b^{2}\right)+\left(c^{4}+4 c^{2}\right)+\left(d^{4}+4 d^{2}\right) \\
& \geq 2\left(\sqrt{a^{4} \cdot 4 a^{2}}+\sqrt{b^{4} \cdot 4 b^{2}}+\sqrt{c^{4} \cdot 4 c^{2}}+\sqrt{d^{4} \cdot 4 d^{2}}\right) \\
& =4\left(\left|a^{3}\right|+\left|b^{3}\right|+\left|c^{3}\right|+\left|d^{3}\right|\right) \geq 4\left(a^{3}+b^{3}+c^{3}+d^{3}\right),
\end{aligned}
$$

which establishes the rightmost inequality.
To prove the leftmost inequality, we first show that $a, b, c, d \in[0,3]$ as in the previous solution. Moreover, we can assume that $0 \leq a \leq b \leq c \leq d$. Then we have $a+b \leq b+c \leq$ $\frac{2}{3}(b+c+d) \leq \frac{2}{3} \cdot 6=4$.

Next, we show that $4 b-b^{2} \leq 4 c-c^{2}$. Actually, this inequality rewrites as $(c-b)(b+c-4) \leq 0$, which follows from the previous estimate. The inequality $4 a-a^{2} \leq 4 b-b^{2}$ can be proved analogously.

Further, the inequalities $a \leq b \leq c$ together with $4 a-a^{2} \leq 4 b-b^{2} \leq 4 c-c^{2}$ allow us to apply the Chebyshev inequality obtaining

$$
\begin{aligned}
a^{2}\left(4 a-a^{2}\right)+b^{2}\left(4 b-b^{2}\right)+c^{2}\left(4 c-c^{2}\right) & \geq \frac{1}{3}\left(a^{2}+b^{2}+c^{2}\right)\left(4(a+b+c)-\left(a^{2}+b^{2}+c^{2}\right)\right) \\
& =\frac{\left(12-d^{2}\right)\left(4(6-d)-\left(12-d^{2}\right)\right)}{3}
\end{aligned}
$$

This implies that

$$
\begin{align*}
\left(4 a^{3}-a^{4}\right) & +\left(4 b^{3}-b^{4}\right)+\left(4 c^{3}-c^{4}\right)+\left(4 d^{3}-d^{4}\right) \geq \frac{\left(12-d^{2}\right)\left(d^{2}-4 d+12\right)}{3}+4 d^{3}-d^{4} \\
& =\frac{144-48 d+16 d^{3}-4 d^{4}}{3}=36+\frac{4}{3}(3-d)(d-1)\left(d^{2}-3\right) \tag{2}
\end{align*}
$$

Finally, we have $d^{2} \geq \frac{1}{4}\left(a^{2}+b^{2}+c^{2}+d^{2}\right)=3$ (which implies $d>1$ ); so, the expression $\frac{4}{3}(3-d)(d-1)\left(d^{2}-3\right)$ in the right-hand part of $(2)$ is nonnegative, and the desired inequality is proved.
Comment. The rightmost inequality is easier than the leftmost one. In particular, Solutions 2 and 3 show that only the condition $a^{2}+b^{2}+c^{2}+d^{2}=12$ is needed for the former one.

A3. Let $x_{1}, \ldots, x_{100}$ be nonnegative real numbers such that $x_{i}+x_{i+1}+x_{i+2} \leq 1$ for all $i=1, \ldots, 100$ (we put $x_{101}=x_{1}, x_{102}=x_{2}$ ). Find the maximal possible value of the sum

$$
S=\sum_{i=1}^{100} x_{i} x_{i+2}
$$

(Russia)
Answer. $\frac{25}{2}$.
Solution 1. Let $x_{2 i}=0, x_{2 i-1}=\frac{1}{2}$ for all $i=1, \ldots, 50$. Then we have $S=50 \cdot\left(\frac{1}{2}\right)^{2}=\frac{25}{2}$. So, we are left to show that $S \leq \frac{25}{2}$ for all values of $x_{i}$ 's satisfying the problem conditions.

Consider any $1 \leq i \leq 50$. By the problem condition, we get $x_{2 i-1} \leq 1-x_{2 i}-x_{2 i+1}$ and $x_{2 i+2} \leq 1-x_{2 i}-x_{2 i+1}$. Hence by the AM-GM inequality we get

$$
\begin{aligned}
x_{2 i-1} x_{2 i+1} & +x_{2 i} x_{2 i+2} \leq\left(1-x_{2 i}-x_{2 i+1}\right) x_{2 i+1}+x_{2 i}\left(1-x_{2 i}-x_{2 i+1}\right) \\
& =\left(x_{2 i}+x_{2 i+1}\right)\left(1-x_{2 i}-x_{2 i+1}\right) \leq\left(\frac{\left(x_{2 i}+x_{2 i+1}\right)+\left(1-x_{2 i}-x_{2 i+1}\right)}{2}\right)^{2}=\frac{1}{4} .
\end{aligned}
$$

Summing up these inequalities for $i=1,2, \ldots, 50$, we get the desired inequality

$$
\sum_{i=1}^{50}\left(x_{2 i-1} x_{2 i+1}+x_{2 i} x_{2 i+2}\right) \leq 50 \cdot \frac{1}{4}=\frac{25}{2} .
$$

Comment. This solution shows that a bit more general fact holds. Namely, consider $2 n$ nonnegative numbers $x_{1}, \ldots, x_{2 n}$ in a row (with no cyclic notation) and suppose that $x_{i}+x_{i+1}+x_{i+2} \leq 1$ for all $i=1,2, \ldots, 2 n-2$. Then $\sum_{i=1}^{2 n-2} x_{i} x_{i+2} \leq \frac{n-1}{4}$.

The proof is the same as above, though if might be easier to find it (for instance, applying induction). The original estimate can be obtained from this version by considering the sequence $x_{1}, x_{2}, \ldots, x_{100}, x_{1}, x_{2}$.

Solution 2. We present another proof of the estimate. From the problem condition, we get

$$
\begin{aligned}
S=\sum_{i=1}^{100} x_{i} x_{i+2} \leq \sum_{i=1}^{100} x_{i}\left(1-x_{i}-x_{i+1}\right) & =\sum_{i=1}^{100} x_{i}-\sum_{i=1}^{100} x_{i}^{2}-\sum_{i=1}^{100} x_{i} x_{i+1} \\
& =\sum_{i=1}^{100} x_{i}-\frac{1}{2} \sum_{i=1}^{100}\left(x_{i}+x_{i+1}\right)^{2} .
\end{aligned}
$$

By the AM-QM inequality, we have $\sum\left(x_{i}+x_{i+1}\right)^{2} \geq \frac{1}{100}\left(\sum\left(x_{i}+x_{i+1}\right)\right)^{2}$, so

$$
\begin{aligned}
S \leq \sum_{i=1}^{100} x_{i}-\frac{1}{200}\left(\sum_{i=1}^{100}\left(x_{i}+x_{i+1}\right)\right)^{2} & =\sum_{i=1}^{100} x_{i}-\frac{2}{100}\left(\sum_{i=1}^{100} x_{i}\right)^{2} \\
& =\frac{2}{100}\left(\sum_{i=1}^{100} x_{i}\right)\left(\frac{100}{2}-\sum_{i=1}^{100} x_{i}\right) .
\end{aligned}
$$

And finally, by the AM-GM inequality

$$
S \leq \frac{2}{100} \cdot\left(\frac{1}{2}\left(\sum_{i=1}^{100} x_{i}+\frac{100}{2}-\sum_{i=1}^{100} x_{i}\right)\right)^{2}=\frac{2}{100} \cdot\left(\frac{100}{4}\right)^{2}=\frac{25}{2}
$$

Comment. These solutions are not as easy as they may seem at the first sight. There are two different optimal configurations in which the variables have different values, and not all of sums of three consecutive numbers equal 1. Although it is easy to find the value $\frac{25}{2}$, the estimates must be done with care to preserve equality in the optimal configurations.

A4. A sequence $x_{1}, x_{2}, \ldots$ is defined by $x_{1}=1$ and $x_{2 k}=-x_{k}, x_{2 k-1}=(-1)^{k+1} x_{k}$ for all $k \geq 1$. Prove that $x_{1}+x_{2}+\cdots+x_{n} \geq 0$ for all $n \geq 1$.
(Austria)
Solution 1. We start with some observations. First, from the definition of $x_{i}$ it follows that for each positive integer $k$ we have

$$
\begin{equation*}
x_{4 k-3}=x_{2 k-1}=-x_{4 k-2} \quad \text { and } \quad x_{4 k-1}=x_{4 k}=-x_{2 k}=x_{k} . \tag{1}
\end{equation*}
$$

Hence, denoting $S_{n}=\sum_{i=1}^{n} x_{i}$, we have

$$
\begin{gather*}
S_{4 k}=\sum_{i=1}^{k}\left(\left(x_{4 k-3}+x_{4 k-2}\right)+\left(x_{4 k-1}+x_{4 k}\right)\right)=\sum_{i=1}^{k}\left(0+2 x_{k}\right)=2 S_{k},  \tag{2}\\
S_{4 k+2}=S_{4 k}+\left(x_{4 k+1}+x_{4 k+2}\right)=S_{4 k} . \tag{3}
\end{gather*}
$$

Observe also that $S_{n}=\sum_{i=1}^{n} x_{i} \equiv \sum_{i=1}^{n} 1=n(\bmod 2)$.
Now we prove by induction on $k$ that $S_{i} \geq 0$ for all $i \leq 4 k$. The base case is valid since $x_{1}=x_{3}=x_{4}=1, x_{2}=-1$. For the induction step, assume that $S_{i} \geq 0$ for all $i \leq 4 k$. Using the relations (1)-(3), we obtain

$$
S_{4 k+4}=2 S_{k+1} \geq 0, \quad S_{4 k+2}=S_{4 k} \geq 0, \quad S_{4 k+3}=S_{4 k+2}+x_{4 k+3}=\frac{S_{4 k+2}+S_{4 k+4}}{2} \geq 0
$$

So, we are left to prove that $S_{4 k+1} \geq 0$. If $k$ is odd, then $S_{4 k}=2 S_{k} \geq 0$; since $k$ is odd, $S_{k}$ is odd as well, so we have $S_{4 k} \geq 2$ and hence $S_{4 k+1}=S_{4 k}+x_{4 k+1} \geq 1$.

Conversely, if $k$ is even, then we have $x_{4 k+1}=x_{2 k+1}=x_{k+1}$, hence $S_{4 k+1}=S_{4 k}+x_{4 k+1}=$ $2 S_{k}+x_{k+1}=S_{k}+S_{k+1} \geq 0$. The step is proved.

Solution 2. We will use the notation of $S_{n}$ and the relations (1)-(3) from the previous solution.

Assume the contrary and consider the minimal $n$ such that $S_{n+1}<0$; surely $n \geq 1$, and from $S_{n} \geq 0$ we get $S_{n}=0, x_{n+1}=-1$. Hence, we are especially interested in the set $M=\left\{n: S_{n}=0\right\}$; our aim is to prove that $x_{n+1}=1$ whenever $n \in M$ thus coming to a contradiction.

For this purpose, we first describe the set $M$ inductively. We claim that (i) $M$ consists only of even numbers, (ii) $2 \in M$, and (iii) for every even $n \geq 4$ we have $n \in M \Longleftrightarrow[n / 4] \in M$. Actually, (i) holds since $S_{n} \equiv n(\bmod 2)$, (ii) is straightforward, while (iii) follows from the relations $S_{4 k+2}=S_{4 k}=2 S_{k}$.

Now, we are left to prove that $x_{n+1}=1$ if $n \in M$. We use the induction on $n$. The base case is $n=2$, that is, the minimal element of $M$; here we have $x_{3}=1$, as desired.

For the induction step, consider some $4 \leq n \in M$ and let $m=[n / 4] \in M$; then $m$ is even, and $x_{m+1}=1$ by the induction hypothesis. We prove that $x_{n+1}=x_{m+1}=1$. If $n=4 m$ then we have $x_{n+1}=x_{2 m+1}=x_{m+1}$ since $m$ is even; otherwise, $n=4 m+2$, and $x_{n+1}=-x_{2 m+2}=x_{m+1}$, as desired. The proof is complete.
Comment. Using the inductive definition of set $M$, one can describe it explicitly. Namely, $M$ consists exactly of all positive integers not containing digits 1 and 3 in their 4 -base representation.

A5. Denote by $\mathbb{Q}^{+}$the set of all positive rational numbers. Determine all functions $f: \mathbb{Q}^{+} \rightarrow \mathbb{Q}^{+}$ which satisfy the following equation for all $x, y \in \mathbb{Q}^{+}$:

$$
\begin{equation*}
f\left(f(x)^{2} y\right)=x^{3} f(x y) \tag{1}
\end{equation*}
$$

(Switzerland)
Answer. The only such function is $f(x)=\frac{1}{x}$.
Solution. By substituting $y=1$, we get

$$
\begin{equation*}
f\left(f(x)^{2}\right)=x^{3} f(x) \tag{2}
\end{equation*}
$$

Then, whenever $f(x)=f(y)$, we have

$$
x^{3}=\frac{f\left(f(x)^{2}\right)}{f(x)}=\frac{f\left(f(y)^{2}\right)}{f(y)}=y^{3}
$$

which implies $x=y$, so the function $f$ is injective.
Now replace $x$ by $x y$ in (2), and apply (1) twice, second time to $\left(y, f(x)^{2}\right)$ instead of $(x, y)$ :

$$
f\left(f(x y)^{2}\right)=(x y)^{3} f(x y)=y^{3} f\left(f(x)^{2} y\right)=f\left(f(x)^{2} f(y)^{2}\right)
$$

Since $f$ is injective, we get

$$
\begin{aligned}
f(x y)^{2} & =f(x)^{2} f(y)^{2} \\
f(x y) & =f(x) f(y)
\end{aligned}
$$

Therefore, $f$ is multiplicative. This also implies $f(1)=1$ and $f\left(x^{n}\right)=f(x)^{n}$ for all integers $n$.
Then the function equation (1) can be re-written as

$$
\begin{align*}
f(f(x))^{2} f(y) & =x^{3} f(x) f(y) \\
f(f(x)) & =\sqrt{x^{3} f(x)} \tag{3}
\end{align*}
$$

Let $g(x)=x f(x)$. Then, by (3), we have

$$
\begin{aligned}
g(g(x)) & =g(x f(x))=x f(x) \cdot f(x f(x))=x f(x)^{2} f(f(x))= \\
& =x f(x)^{2} \sqrt{x^{3} f(x)}=(x f(x))^{5 / 2}=(g(x))^{5 / 2}
\end{aligned}
$$

and, by induction,
for every positive integer $n$.
Consider (4) for a fixed $x$. The left-hand side is always rational, so $(g(x))^{(5 / 2)^{n}}$ must be rational for every $n$. We show that this is possible only if $g(x)=1$. Suppose that $g(x) \neq 1$, and let the prime factorization of $g(x)$ be $g(x)=p_{1}^{\alpha_{1}} \ldots p_{k}^{\alpha_{k}}$ where $p_{1}, \ldots, p_{k}$ are distinct primes and $\alpha_{1}, \ldots, \alpha_{k}$ are nonzero integers. Then the unique prime factorization of (4) is

$$
\underbrace{g(g(\ldots g}_{n+1}(x) \ldots))=(g(x))^{(5 / 2)^{n}}=p_{1}^{(5 / 2)^{n} \alpha_{1}} \cdots p_{k}^{(5 / 2)^{n} \alpha_{k}}
$$

where the exponents should be integers. But this is not true for large values of $n$, for example $\left(\frac{5}{2}\right)^{n} \alpha_{1}$ cannot be a integer number when $2^{n} \nmid \alpha_{1}$. Therefore, $g(x) \neq 1$ is impossible.

Hence, $g(x)=1$ and thus $f(x)=\frac{1}{x}$ for all $x$.
The function $f(x)=\frac{1}{x}$ satisfies the equation (1):

$$
f\left(f(x)^{2} y\right)=\frac{1}{f(x)^{2} y}=\frac{1}{\left(\frac{1}{x}\right)^{2} y}=\frac{x^{3}}{x y}=x^{3} f(x y)
$$

Comment. Among $\mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$functions, $f(x)=\frac{1}{x}$ is not the only solution. Another solution is $f_{1}(x)=x^{3 / 2}$. Using transfinite tools, infinitely many other solutions can be constructed.

A6. Suppose that $f$ and $g$ are two functions defined on the set of positive integers and taking positive integer values. Suppose also that the equations $f(g(n))=f(n)+1$ and $g(f(n))=$ $g(n)+1$ hold for all positive integers. Prove that $f(n)=g(n)$ for all positive integer $n$.
(Germany)
Solution 1. Throughout the solution, by $\mathbb{N}$ we denote the set of all positive integers. For any function $h: \mathbb{N} \rightarrow \mathbb{N}$ and for any positive integer $k$, define $h^{k}(x)=\underbrace{h(h(\ldots h}_{k}(x) \ldots)$ ) (in particular, $\left.h^{0}(x)=x\right)$.

Observe that $f\left(g^{k}(x)\right)=f\left(g^{k-1}(x)\right)+1=\cdots=f(x)+k$ for any positive integer $k$, and similarly $g\left(f^{k}(x)\right)=g(x)+k$. Now let $a$ and $b$ are the minimal values attained by $f$ and $g$, respectively; say $f\left(n_{f}\right)=a, g\left(n_{g}\right)=b$. Then we have $f\left(g^{k}\left(n_{f}\right)\right)=a+k, g\left(f^{k}\left(n_{g}\right)\right)=b+k$, so the function $f$ attains all values from the set $N_{f}=\{a, a+1, \ldots\}$, while $g$ attains all the values from the set $N_{g}=\{b, b+1, \ldots\}$.

Next, note that $f(x)=f(y)$ implies $g(x)=g(f(x))-1=g(f(y))-1=g(y)$; surely, the converse implication also holds. Now, we say that $x$ and $y$ are similar (and write $x \sim y$ ) if $f(x)=f(y)$ (equivalently, $g(x)=g(y)$ ). For every $x \in \mathbb{N}$, we define $[x]=\{y \in \mathbb{N}: x \sim y\}$; surely, $y_{1} \sim y_{2}$ for all $y_{1}, y_{2} \in[x]$, so $[x]=[y]$ whenever $y \in[x]$.

Now we investigate the structure of the sets $[x]$.
Claim 1. Suppose that $f(x) \sim f(y)$; then $x \sim y$, that is, $f(x)=f(y)$. Consequently, each class [ $x$ ] contains at most one element from $N_{f}$, as well as at most one element from $N_{g}$.
Proof. If $f(x) \sim f(y)$, then we have $g(x)=g(f(x))-1=g(f(y))-1=g(y)$, so $x \sim y$. The second statement follows now from the sets of values of $f$ and $g$.

Next, we clarify which classes do not contain large elements.
Claim 2. For any $x \in \mathbb{N}$, we have $[x] \subseteq\{1,2, \ldots, b-1\}$ if and only if $f(x)=a$. Analogously, $[x] \subseteq\{1,2, \ldots, a-1\}$ if and only if $g(x)=b$.
Proof. We will prove that $[x] \nsubseteq\{1,2, \ldots, b-1\} \Longleftrightarrow f(x)>a$; the proof of the second statement is similar.

Note that $f(x)>a$ implies that there exists some $y$ satisfying $f(y)=f(x)-1$, so $f(g(y))=$ $f(y)+1=f(x)$, and hence $x \sim g(y) \geq b$. Conversely, if $b \leq c \sim x$ then $c=g(y)$ for some $y \in \mathbb{N}$, which in turn follows $f(x)=f(g(y))=f(y)+1 \geq a+1$, and hence $f(x)>a$.

Claim 2 implies that there exists exactly one class contained in $\{1, \ldots, a-1\}$ (that is, the class $\left[n_{g}\right]$ ), as well as exactly one class contained in $\{1, \ldots, b-1\}$ (the class $\left[n_{f}\right]$ ). Assume for a moment that $a \leq b$; then $\left[n_{g}\right]$ is contained in $\{1, \ldots, b-1\}$ as well, hence it coincides with $\left[n_{g}\right]$. So, we get that

$$
\begin{equation*}
f(x)=a \Longleftrightarrow g(x)=b \Longleftrightarrow x \sim n_{f} \sim n_{g} . \tag{1}
\end{equation*}
$$

Claim 3. $a=b$.
Proof. By Claim 2, we have $[a] \neq\left[n_{f}\right]$, so $[a]$ should contain some element $a^{\prime} \geq b$ by Claim 2 again. If $a \neq a^{\prime}$, then $[a]$ contains two elements $\geq a$ which is impossible by Claim 1 . Therefore, $a=a^{\prime} \geq b$. Similarly, $b \geq a$.

Now we are ready to prove the problem statement. First, we establish the following
Claim 4. For every integer $d \geq 0, f^{d+1}\left(n_{f}\right)=g^{d+1}\left(n_{f}\right)=a+d$.
Proof. Induction on $d$. For $d=0$, the statement follows from (1) and Claim 3. Next, for $d>1$ from the induction hypothesis we have $f^{d+1}\left(n_{f}\right)=f\left(f^{d}\left(n_{f}\right)\right)=f\left(g^{d}\left(n_{f}\right)\right)=f\left(n_{f}\right)+d=a+d$. The equality $g^{d+1}\left(n_{f}\right)=a+d$ is analogous.

Finally, for each $x \in \mathbb{N}$, we have $f(x)=a+d$ for some $d \geq 0$, so $f(x)=f\left(g^{d}\left(n_{f}\right)\right)$ and hence $x \sim g^{d}\left(n_{f}\right)$. It follows that $g(x)=g\left(g^{d}\left(n_{f}\right)\right)=g^{d+1}\left(n_{f}\right)=a+d=f(x)$ by Claim 4 .

Solution 2. We start with the same observations, introducing the relation $\sim$ and proving Claim 1 from the previous solution.

Note that $f(a)>a$ since otherwise we have $f(a)=a$ and hence $g(a)=g(f(a))=g(a)+1$, which is false.
Claim 2'. $a=b$.
Proof. We can assume that $a \leq b$. Since $f(a) \geq a+1$, there exists some $x \in \mathbb{N}$ such that $f(a)=f(x)+1$, which is equivalent to $f(a)=f(g(x))$ and $a \sim g(x)$. Since $g(x) \geq b \geq a$, by Claim 1 we have $a=g(x) \geq b$, which together with $a \leq b$ proves the Claim.

Now, almost the same method allows to find the values $f(a)$ and $g(a)$.
Claim 3'. $f(a)=g(a)=a+1$.
Proof. Assume the contrary; then $f(a) \geq a+2$, hence there exist some $x, y \in \mathbb{N}$ such that $f(x)=f(a)-2$ and $f(y)=g(x)($ as $g(x) \geq a=b)$. Now we get $f(a)=f(x)+2=f\left(g^{2}(x)\right)$, so $a \sim g^{2}(x) \geq a$, and by Claim 1 we get $a=g^{2}(x)=g(f(y))=1+g(y) \geq 1+a$; this is impossible. The equality $g(a)=a+1$ is similar.

Now, we are prepared for the proof of the problem statement. First, we prove it for $n \geq a$. Claim 4'. For each integer $x \geq a$, we have $f(x)=g(x)=x+1$.
Proof. Induction on $x$. The base case $x=a$ is provided by Claim $3^{\prime}$, while the induction step follows from $f(x+1)=f(g(x))=f(x)+1=(x+1)+1$ and the similar computation for $g(x+1)$.

Finally, for an arbitrary $n \in \mathbb{N}$ we have $g(n) \geq a$, so by Claim $4^{\prime}$ we have $f(n)+1=$ $f(g(n))=g(n)+1$, hence $f(n)=g(n)$.
Comment. It is not hard now to describe all the functions $f: \mathbb{N} \rightarrow \mathbb{N}$ satisfying the property $f(f(n))=$ $f(n)+1$. For each such function, there exists $n_{0} \in \mathbb{N}$ such that $f(n)=n+1$ for all $n \geq n_{0}$, while for each $n<n_{0}, f(n)$ is an arbitrary number greater than of equal to $n_{0}$ (these numbers may be different for different $n<n_{0}$ ).

A7. Let $a_{1}, \ldots, a_{r}$ be positive real numbers. For $n>r$, we inductively define

$$
\begin{equation*}
a_{n}=\max _{1 \leq k \leq n-1}\left(a_{k}+a_{n-k}\right) \tag{1}
\end{equation*}
$$

Prove that there exist positive integers $\ell \leq r$ and $N$ such that $a_{n}=a_{n-\ell}+a_{\ell}$ for all $n \geq N$.

Solution 1. First, from the problem conditions we have that each $a_{n}(n>r)$ can be expressed as $a_{n}=a_{j_{1}}+a_{j_{2}}$ with $j_{1}, j_{2}<n, j_{1}+j_{2}=n$. If, say, $j_{1}>r$ then we can proceed in the same way with $a_{j_{1}}$, and so on. Finally, we represent $a_{n}$ in a form

$$
\begin{gather*}
a_{n}=a_{i_{1}}+\cdots+a_{i_{k}},  \tag{2}\\
1 \leq i_{j} \leq r, \quad i_{1}+\cdots+i_{k}=n . \tag{3}
\end{gather*}
$$

Moreover, if $a_{i_{1}}$ and $a_{i_{2}}$ are the numbers in (2) obtained on the last step, then $i_{1}+i_{2}>r$. Hence we can adjust (3) as

$$
\begin{equation*}
1 \leq i_{j} \leq r, \quad i_{1}+\cdots+i_{k}=n, \quad i_{1}+i_{2}>r . \tag{4}
\end{equation*}
$$

On the other hand, suppose that the indices $i_{1}, \ldots, i_{k}$ satisfy the conditions (4). Then, denoting $s_{j}=i_{1}+\cdots+i_{j}$, from (1) we have

$$
a_{n}=a_{s_{k}} \geq a_{s_{k-1}}+a_{i_{k}} \geq a_{s_{k-2}}+a_{i_{k-1}}+a_{i_{k}} \geq \cdots \geq a_{i_{1}}+\cdots+a_{i_{k}} .
$$

Summarizing these observations we get the following
Claim. For every $n>r$, we have

$$
a_{n}=\max \left\{a_{i_{1}}+\cdots+a_{i_{k}}: \text { the collection }\left(i_{1}, \ldots, i_{k}\right) \text { satisfies }(4)\right\} .
$$

Now we denote

$$
s=\max _{1 \leq i \leq r} \frac{a_{i}}{i}
$$

and fix some index $\ell \leq r$ such that $s=\frac{a_{\ell}}{\ell}$.
Consider some $n \geq r^{2} \ell+2 r$ and choose an expansion of $a_{n}$ in the form (2), (4). Then we have $n=i_{1}+\cdots+i_{k} \leq r k$, so $k \geq n / r \geq r \ell+2$. Suppose that none of the numbers $i_{3}, \ldots, i_{k}$ equals $\ell$. Then by the pigeonhole principle there is an index $1 \leq j \leq r$ which appears among $i_{3}, \ldots, i_{k}$ at least $\ell$ times, and surely $j \neq \ell$. Let us delete these $\ell$ occurrences of $j$ from $\left(i_{1}, \ldots, i_{k}\right)$, and add $j$ occurrences of $\ell$ instead, obtaining a sequence $\left(i_{1}, i_{2}, i_{3}^{\prime}, \ldots, i_{k^{\prime}}^{\prime}\right)$ also satisfying (4). By Claim, we have

$$
a_{i_{1}}+\cdots+a_{i_{k}}=a_{n} \geq a_{i_{1}}+a_{i_{2}}+a_{i_{3}^{\prime}}+\cdots+a_{i_{k_{k}^{\prime}}^{\prime}}
$$

or, after removing the coinciding terms, $\ell a_{j} \geq j a_{\ell}$, so $\frac{a_{\ell}}{\ell} \leq \frac{a_{j}}{j}$. By the definition of $\ell$, this means that $\ell a_{j}=j a_{\ell}$, hence

$$
a_{n}=a_{i_{1}}+a_{i_{2}}+a_{i_{3}^{\prime}}+\cdots+a_{i_{k^{\prime}}^{\prime}}
$$

Thus, for every $n \geq r^{2} \ell+2 r$ we have found a representation of the form (2), (4) with $i_{j}=\ell$ for some $j \geq 3$. Rearranging the indices we may assume that $i_{k}=\ell$.

Finally, observe that in this representation, the indices $\left(i_{1}, \ldots, i_{k-1}\right)$ satisfy the conditions (4) with $n$ replaced by $n-\ell$. Thus, from the Claim we get

$$
a_{n-\ell}+a_{\ell} \geq\left(a_{i_{1}}+\cdots+a_{i_{k-1}}\right)+a_{\ell}=a_{n}
$$

which by (1) implies

$$
a_{n}=a_{n-\ell}+a_{\ell} \quad \text { for each } n \geq r^{2} \ell+2 r,
$$

as desired.

Solution 2. As in the previous solution, we involve the expansion (2), (3), and we fix some index $1 \leq \ell \leq r$ such that

$$
\frac{a_{\ell}}{\ell}=s=\max _{1 \leq i \leq r} \frac{a_{i}}{i}
$$

Now, we introduce the sequence $\left(b_{n}\right)$ as $b_{n}=a_{n}-s n$; then $b_{\ell}=0$.
We prove by induction on $n$ that $b_{n} \leq 0$, and $\left(b_{n}\right)$ satisfies the same recurrence relation as $\left(a_{n}\right)$. The base cases $n \leq r$ follow from the definition of $s$. Now, for $n>r$ from the induction hypothesis we have

$$
b_{n}=\max _{1 \leq k \leq n-1}\left(a_{k}+a_{n-k}\right)-n s=\max _{1 \leq k \leq n-1}\left(b_{k}+b_{n-k}+n s\right)-n s=\max _{1 \leq k \leq n-1}\left(b_{k}+b_{n-k}\right) \leq 0,
$$

as required.
Now, if $b_{k}=0$ for all $1 \leq k \leq r$, then $b_{n}=0$ for all $n$, hence $a_{n}=s n$, and the statement is trivial. Otherwise, define

$$
M=\max _{1 \leq i \leq r}\left|b_{i}\right|, \quad \varepsilon=\min \left\{\left|b_{i}\right|: 1 \leq i \leq r, b_{i}<0\right\} .
$$

Then for $n>r$ we obtain

$$
b_{n}=\max _{1 \leq k \leq n-1}\left(b_{k}+b_{n-k}\right) \geq b_{\ell}+b_{n-\ell}=b_{n-\ell}
$$

so

$$
0 \geq b_{n} \geq b_{n-\ell} \geq b_{n-2 \ell} \geq \cdots \geq-M
$$

Thus, in view of the expansion (2), (3) applied to the sequence $\left(b_{n}\right)$, we get that each $b_{n}$ is contained in a set

$$
T=\left\{b_{i_{1}}+b_{i_{2}}+\cdots+b_{i_{k}}: i_{1}, \ldots, i_{k} \leq r\right\} \cap[-M, 0]
$$

We claim that this set is finite. Actually, for any $x \in T$, let $x=b_{i_{1}}+\cdots+b_{i_{k}}\left(i_{1}, \ldots, i_{k} \leq r\right)$. Then among $b_{i_{j}}$ 's there are at most $\frac{M}{\varepsilon}$ nonzero terms (otherwise $x<\frac{M}{\varepsilon} \cdot(-\varepsilon)<-M$ ). Thus $x$ can be expressed in the same way with $k \leq \frac{M}{\varepsilon}$, and there is only a finite number of such sums.

Finally, for every $t=1,2, \ldots, \ell$ we get that the sequence

$$
b_{r+t}, b_{r+t+\ell}, b_{r+t+2 \ell}, \ldots
$$

is non-decreasing and attains the finite number of values; therefore it is constant from some index. Thus, the sequence $\left(b_{n}\right)$ is periodic with period $\ell$ from some index $N$, which means that

$$
b_{n}=b_{n-\ell}=b_{n-\ell}+b_{\ell} \quad \text { for all } n>N+\ell
$$

and hence

$$
a_{n}=b_{n}+n s=\left(b_{n-\ell}+(n-\ell) s\right)+\left(b_{\ell}+\ell s\right)=a_{n-\ell}+a_{\ell} \quad \text { for all } n>N+\ell,
$$

as desired.

A8. Given six positive numbers $a, b, c, d, e, f$ such that $a<b<c<d<e<f$. Let $a+c+e=S$ and $b+d+f=T$. Prove that

$$
\begin{equation*}
2 S T>\sqrt{3(S+T)(S(b d+b f+d f)+T(a c+a e+c e))} . \tag{1}
\end{equation*}
$$

(South Korea)
Solution 1. We define also $\sigma=a c+c e+a e, \tau=b d+b f+d f$. The idea of the solution is to interpret (1) as a natural inequality on the roots of an appropriate polynomial.

Actually, consider the polynomial

$$
\begin{align*}
& P(x)=(b+d+f)(x-a)(x-c)(x-e)+(a+c+e)(x-b)(x-d)(x-f) \\
&=T\left(x^{3}-S x^{2}+\sigma x-a c e\right)+S\left(x^{3}-T x^{2}+\tau x-b d f\right) \tag{2}
\end{align*}
$$

Surely, $P$ is cubic with leading coefficient $S+T>0$. Moreover, we have

$$
\begin{array}{ll}
P(a)=S(a-b)(a-d)(a-f)<0, & P(c)=S(c-b)(c-d)(c-f)>0 \\
P(e)=S(e-b)(e-d)(e-f)<0, & P(f)=T(f-a)(f-c)(f-e)>0 .
\end{array}
$$

Hence, each of the intervals $(a, c),(c, e),(e, f)$ contains at least one root of $P(x)$. Since there are at most three roots at all, we obtain that there is exactly one root in each interval (denote them by $\alpha \in(a, c), \beta \in(c, e), \gamma \in(e, f))$. Moreover, the polynomial $P$ can be factorized as

$$
\begin{equation*}
P(x)=(T+S)(x-\alpha)(x-\beta)(x-\gamma) \tag{3}
\end{equation*}
$$

Equating the coefficients in the two representations (2) and (3) of $P(x)$ provides

$$
\alpha+\beta+\gamma=\frac{2 T S}{T+S}, \quad \alpha \beta+\alpha \gamma+\beta \gamma=\frac{S \tau+T \sigma}{T+S}
$$

Now, since the numbers $\alpha, \beta, \gamma$ are distinct, we have

$$
0<(\alpha-\beta)^{2}+(\alpha-\gamma)^{2}+(\beta-\gamma)^{2}=2(\alpha+\beta+\gamma)^{2}-6(\alpha \beta+\alpha \gamma+\beta \gamma)
$$

which implies

$$
\frac{4 S^{2} T^{2}}{(T+S)^{2}}=(\alpha+\beta+\gamma)^{2}>3(\alpha \beta+\alpha \gamma+\beta \gamma)=\frac{3(S \tau+T \sigma)}{T+S}
$$

or

$$
4 S^{2} T^{2}>3(T+S)(T \sigma+S \tau)
$$

which is exactly what we need.
Comment 1. In fact, one can locate the roots of $P(x)$ more narrowly: they should lie in the intervals $(a, b),(c, d),(e, f)$.

Surely, if we change all inequality signs in the problem statement to non-strict ones, the (non-strict) inequality will also hold by continuity. One can also find when the equality is achieved. This happens in that case when $P(x)$ is a perfect cube, which immediately implies that $b=c=d=e(=\alpha=\beta=\gamma)$, together with the additional condition that $P^{\prime \prime}(b)=0$. Algebraically,

$$
\begin{array}{rlr}
6(T+S) b-4 T S=0 & \Longleftrightarrow & 3 b(a+4 b+f)=2(a+2 b)(2 b+f) \\
& \Longleftrightarrow & f=\frac{b(4 b-a)}{2 a+b}=b\left(1+\frac{3(b-a)}{2 a+b}\right)>b .
\end{array}
$$

This means that for every pair of numbers $a, b$ such that $0<a<b$, there exists $f>b$ such that the point $(a, b, b, b, b, f)$ is a point of equality.

Solution 2. Let

$$
U=\frac{1}{2}\left((e-a)^{2}+(c-a)^{2}+(e-c)^{2}\right)=S^{2}-3(a c+a e+c e)
$$

and

$$
V=\frac{1}{2}\left((f-b)^{2}+(f-d)^{2}+(d-b)^{2}\right)=T^{2}-3(b d+b f+d f) .
$$

Then

$$
\begin{aligned}
& \text { (L.H.S.) })^{2}-(\text { R.H.S. })^{2}=(2 S T)^{2}-(S+T)(S \cdot 3(b d+b f+d f)+T \cdot 3(a c+a e+c e))= \\
& \quad=4 S^{2} T^{2}-(S+T)\left(S\left(T^{2}-V\right)+T\left(S^{2}-U\right)\right)=(S+T)(S V+T U)-S T(T-S)^{2},
\end{aligned}
$$

and the statement is equivalent with

$$
\begin{equation*}
(S+T)(S V+T U)>S T(T-S)^{2} \tag{4}
\end{equation*}
$$

By the Cauchy-Schwarz inequality,

$$
\begin{equation*}
(S+T)(T U+S V) \geq(\sqrt{S \cdot T U}+\sqrt{T \cdot S V})^{2}=S T(\sqrt{U}+\sqrt{V})^{2} . \tag{5}
\end{equation*}
$$

Estimate the quantities $\sqrt{U}$ and $\sqrt{V}$ by the QM-AM inequality with the positive terms $(e-c)^{2}$ and $(d-b)^{2}$ being omitted:

$$
\begin{align*}
\sqrt{U}+\sqrt{V} & >\sqrt{\frac{(e-a)^{2}+(c-a)^{2}}{2}}+\sqrt{\frac{(f-b)^{2}+(f-d)^{2}}{2}} \\
& >\frac{(e-a)+(c-a)}{2}+\frac{(f-b)+(f-d)}{2}=\left(f-\frac{d}{2}-\frac{b}{2}\right)+\left(\frac{e}{2}+\frac{c}{2}-a\right) \\
& =(T-S)+\frac{3}{2}(e-d)+\frac{3}{2}(c-b)>T-S . \tag{6}
\end{align*}
$$

The estimates (5) and (6) prove (4) and hence the statement.
Solution 3. We keep using the notations $\sigma$ and $\tau$ from Solution 1. Moreover, let $s=c+e$. Note that

$$
(c-b)(c-d)+(e-f)(e-d)+(e-f)(c-b)<0
$$

since each summand is negative. This rewrites as

$$
\begin{align*}
(b d+b f+d f)-(a c+c e+a e) & <(c+e)(b+d+f-a-c-e), \text { or } \\
\tau-\sigma & <s(T-S) . \tag{7}
\end{align*}
$$

Then we have

$$
\begin{aligned}
S \tau+T \sigma & =S(\tau-\sigma)+(S+T) \sigma<S s(T-S)+(S+T)(c e+a s) \\
& \leq S s(T-S)+(S+T)\left(\frac{s^{2}}{4}+(S-s) s\right)=s\left(2 S T-\frac{3}{4}(S+T) s\right) .
\end{aligned}
$$

Using this inequality together with the AM-GM inequality we get

$$
\begin{aligned}
\sqrt{\frac{3}{4}(S+T)(S \tau+T \sigma)} & <\sqrt{\frac{3}{4}(S+T) s\left(2 S T-\frac{3}{4}(S+T) s\right)} \\
& \leq \frac{\frac{3}{4}(S+T) s+2 S T-\frac{3}{4}(S+T) s}{2}=S T .
\end{aligned}
$$

Hence,

$$
2 S T>\sqrt{3(S+T)(S(b d+b f+d f)+T(a c+a e+c e))}
$$

Comment 2. The expression (7) can be found by considering the sum of the roots of the quadratic polynomial $q(x)=(x-b)(x-d)(x-f)-(x-a)(x-c)(x-e)$.

Solution 4. We introduce the expressions $\sigma$ and $\tau$ as in the previous solutions. The idea of the solution is to change the values of variables $a, \ldots, f$ keeping the left-hand side unchanged and increasing the right-hand side; it will lead to a simpler inequality which can be proved in a direct way.

Namely, we change the variables (i) keeping the (non-strict) inequalities $a \leq b \leq c \leq d \leq$ $e \leq f$; (ii) keeping the values of sums $S$ and $T$ unchanged; and finally (iii) increasing the values of $\sigma$ and $\tau$. Then the left-hand side of (1) remains unchanged, while the right-hand side increases. Hence, the inequality (1) (and even a non-strict version of (1)) for the changed values would imply the same (strict) inequality for the original values.

First, we find the sufficient conditions for (ii) and (iii) to be satisfied.
Lemma. Let $x, y, z>0$; denote $U(x, y, z)=x+y+z, v(x, y, z)=x y+x z+y z$. Suppose that $x^{\prime}+y^{\prime}=x+y$ but $|x-y| \geq\left|x^{\prime}-y^{\prime}\right| ;$ then we have $U\left(x^{\prime}, y^{\prime}, z\right)=U(x, y, z)$ and $v\left(x^{\prime}, y^{\prime}, z\right) \geq$ $v(x, y, z)$ with equality achieved only when $|x-y|=\left|x^{\prime}-y^{\prime}\right|$.
Proof. The first equality is obvious. For the second, we have

$$
\begin{aligned}
v\left(x^{\prime}, y^{\prime}, z\right)=z\left(x^{\prime}+y^{\prime}\right)+x^{\prime} y^{\prime} & =z\left(x^{\prime}+y^{\prime}\right)+\frac{\left(x^{\prime}+y^{\prime}\right)^{2}-\left(x^{\prime}-y^{\prime}\right)^{2}}{4} \\
& \geq z(x+y)+\frac{(x+y)^{2}-(x-y)^{2}}{4}=v(x, y, z)
\end{aligned}
$$

with the equality achieved only for $\left(x^{\prime}-y^{\prime}\right)^{2}=(x-y)^{2} \Longleftrightarrow\left|x^{\prime}-y^{\prime}\right|=|x-y|$, as desired.

Now, we apply Lemma several times making the following changes. For each change, we denote the new values by the same letters to avoid cumbersome notations.

1. Let $k=\frac{d-c}{2}$. Replace $(b, c, d, e)$ by $(b+k, c+k, d-k, e-k)$. After the change we have $a<b<c=d<e<f$, the values of $S, T$ remain unchanged, but $\sigma, \tau$ strictly increase by Lemma.
2. Let $\ell=\frac{e-d}{2}$. Replace $(c, d, e, f)$ by $(c+\ell, d+\ell, e-\ell, f-\ell)$. After the change we have $a<b<c=d=e<f$, the values of $S, T$ remain unchanged, but $\sigma, \tau$ strictly increase by the Lemma.
3. Finally, let $m=\frac{c-b}{3}$. Replace $(a, b, c, d, e, f)$ by $(a+2 m, b+2 m, c-m, d-m, e-m, f-m)$. After the change, we have $a<b=c=d=e<f$ and $S, T$ are unchanged. To check (iii), we observe that our change can be considered as a composition of two changes: $(a, b, c, d) \rightarrow$ $(a+m, b+m, c-m, d-m)$ and $(a, b, e, f) \rightarrow(a+m, b+m, e-m, f-m)$. It is easy to see that each of these two consecutive changes satisfy the conditions of the Lemma, hence the values of $\sigma$ and $\tau$ increase.

Finally, we come to the situation when $a<b=c=d=e<f$, and we need to prove the inequality

$$
\begin{align*}
2(a+2 b)(2 b+f) & \geq \sqrt{3(a+4 b+f)\left((a+2 b)\left(b^{2}+2 b f\right)+(2 b+f)\left(2 a b+b^{2}\right)\right)} \\
& =\sqrt{3 b(a+4 b+f) \cdot((a+2 b)(b+2 f)+(2 b+f)(2 a+b))} \tag{8}
\end{align*}
$$

Now, observe that

$$
2 \cdot 2(a+2 b)(2 b+f)=3 b(a+4 b+f)+((a+2 b)(b+2 f)+(2 a+b)(2 b+f))
$$

Hence (4) rewrites as

$$
\begin{aligned}
3 b(a+4 b+f) & +((a+2 b)(b+2 f)+(2 a+b)(2 b+f)) \\
& \geq 2 \sqrt{3 b(a+4 b+f) \cdot((a+2 b)(b+2 f)+(2 b+f)(2 a+b))}
\end{aligned}
$$

which is simply the AM-GM inequality.
Comment 3. Here, we also can find all the cases of equality. Actually, it is easy to see that if some two numbers among $b, c, d, e$ are distinct then one can use Lemma to increase the right-hand side of (1). Further, if $b=c=d=e$, then we need equality in (4); this means that we apply AM-GM to equal numbers, that is,

$$
3 b(a+4 b+f)=(a+2 b)(b+2 f)+(2 a+b)(2 b+f),
$$

which leads to the same equality as in Comment 1.

## Combinatorics

C1. In a concert, 20 singers will perform. For each singer, there is a (possibly empty) set of other singers such that he wishes to perform later than all the singers from that set. Can it happen that there are exactly 2010 orders of the singers such that all their wishes are satisfied?
(Austria)
Answer. Yes, such an example exists.
Solution. We say that an order of singers is good if it satisfied all their wishes. Next, we say that a number $N$ is realizable by $k$ singers (or $k$-realizable) if for some set of wishes of these singers there are exactly $N$ good orders. Thus, we have to prove that a number 2010 is 20-realizable.

We start with the following simple
Lemma. Suppose that numbers $n_{1}, n_{2}$ are realizable by respectively $k_{1}$ and $k_{2}$ singers. Then the number $n_{1} n_{2}$ is $\left(k_{1}+k_{2}\right)$-realizable.
Proof. Let the singers $A_{1}, \ldots, A_{k_{1}}$ (with some wishes among them) realize $n_{1}$, and the singers $B_{1}$, $\ldots, B_{k_{2}}$ (with some wishes among them) realize $n_{2}$. Add to each singer $B_{i}$ the wish to perform later than all the singers $A_{j}$. Then, each good order of the obtained set of singers has the form $\left(A_{i_{1}}, \ldots, A_{i_{k_{1}}}, B_{j_{1}}, \ldots, B_{j_{k_{2}}}\right)$, where $\left(A_{i_{1}}, \ldots, A_{i_{k_{1}}}\right)$ is a good order for $A_{i}$ 's and $\left(B_{j_{1}}, \ldots, B_{j_{k_{2}}}\right)$ is a good order for $B_{j}$ 's. Conversely, each order of this form is obviously good. Hence, the number of good orders is $n_{1} n_{2}$.

In view of Lemma, we show how to construct sets of singers containing 4, 3 and 13 singers and realizing the numbers 5,6 and 67 , respectively. Thus the number $2010=6 \cdot 5 \cdot 67$ will be realizable by $4+3+13=20$ singers. These companies of singers are shown in Figs. 1-3; the wishes are denoted by arrows, and the number of good orders for each Figure stands below in the brackets.

(5)

Fig. 1

(67)

Fig. 3

For Fig. 1, there are exactly 5 good orders $(a, b, c, d),(a, b, d, c),(b, a, c, d),(b, a, d, c)$, $(b, d, a, c)$. For Fig. 2, each of 6 orders is good since there are no wishes.

Finally, for Fig. 3, the order of $a_{1}, \ldots, a_{11}$ is fixed; in this line, singer $x$ can stand before each of $a_{i}(i \leq 9)$, and singer $y$ can stand after each of $a_{j}(j \geq 5)$, thus resulting in $9 \cdot 7=63$ cases. Further, the positions of $x$ and $y$ in this line determine the whole order uniquely unless both of them come between the same pair ( $a_{i}, a_{i+1}$ ) (thus $5 \leq i \leq 8$ ); in the latter cases, there are two orders instead of 1 due to the order of $x$ and $y$. Hence, the total number of good orders is $63+4=67$, as desired.

Comment. The number 20 in the problem statement is not sharp and is put there to respect the original formulation. So, if necessary, the difficulty level of this problem may be adjusted by replacing 20 by a smaller number. Here we present some improvements of the example leading to a smaller number of singers.

Surely, each example with $<20$ singers can be filled with some "super-stars" who should perform at the very end in a fixed order. Hence each of these improvements provides a different solution for the problem. Moreover, the large variety of ideas standing behind these examples allows to suggest that there are many other examples.

1. Instead of building the examples realizing 5 and 6 , it is more economic to make an example realizing 30 ; it may seem even simpler. Two possible examples consisting of 5 and 6 singers are shown in Fig. 4; hence the number 20 can be decreased to 19 or 18 .

For Fig. 4a, the order of $a_{1}, \ldots, a_{4}$ is fixed, there are 5 ways to add $x$ into this order, and there are 6 ways to add $y$ into the resulting order of $a_{1}, \ldots, a_{4}, x$. Hence there are $5 \cdot 6=30$ good orders.

On Fig. 4b, for 5 singers $a, b_{1}, b_{2}, c_{1}, c_{2}$ there are $5!=120$ orders at all. Obviously, exactly one half of them satisfies the wish $b_{1} \leftarrow b_{2}$, and exactly one half of these orders satisfies another wish $c_{1} \leftarrow c_{2}$; hence, there are exactly $5!/ 4=30$ good orders.


Fig. 4

(2010)

Fig. 5

(2010)

Fig. 6
2. One can merge the examples for 30 and 67 shown in Figs. 4 b and 3 in a smarter way, obtaining a set of 13 singers representing 2010. This example is shown in Fig. 5; an arrow from/to group $\left\{b_{1}, \ldots, b_{5}\right\}$ means that there exists such arrow from each member of this group.

Here, as in Fig. 4b, one can see that there are exactly 30 orders of $b_{1}, \ldots, b_{5}, a_{6}, \ldots, a_{11}$ satisfying all their wishes among themselves. Moreover, one can prove in the same way as for Fig. 3 that each of these orders can be complemented by $x$ and $y$ in exactly 67 ways, hence obtaining $30 \cdot 67=2010$ good orders at all.

Analogously, one can merge the examples in Figs. 1-3 to represent 2010 by 13 singers as is shown in Fig. 6.


Fig. 7
3. Finally, we will present two other improvements; the proofs are left to the reader. The graph in Fig. 7 shows how 10 singers can represent 67 . Moreover, even a company of 10 singers representing 2010 can be found; this company is shown in Fig. 8.

C2. On some planet, there are $2^{N}$ countries $(N \geq 4)$. Each country has a flag $N$ units wide and one unit high composed of $N$ fields of size $1 \times 1$, each field being either yellow or blue. No two countries have the same flag.

We say that a set of $N$ flags is diverse if these flags can be arranged into an $N \times N$ square so that all $N$ fields on its main diagonal will have the same color. Determine the smallest positive integer $M$ such that among any $M$ distinct flags, there exist $N$ flags forming a diverse set.
(Croatia)
Answer. $M=2^{N-2}+1$.
Solution. When speaking about the diagonal of a square, we will always mean the main diagonal.

Let $M_{N}$ be the smallest positive integer satisfying the problem condition. First, we show that $M_{N}>2^{N-2}$. Consider the collection of all $2^{N-2}$ flags having yellow first squares and blue second ones. Obviously, both colors appear on the diagonal of each $N \times N$ square formed by these flags.

We are left to show that $M_{N} \leq 2^{N-2}+1$, thus obtaining the desired answer. We start with establishing this statement for $N=4$.

Suppose that we have 5 flags of length 4 . We decompose each flag into two parts of 2 squares each; thereby, we denote each flag as $L R$, where the $2 \times 1$ flags $L, R \in \mathcal{S}=\{\mathrm{BB}, \mathrm{BY}, \mathrm{YB}, \mathrm{YY}\}$ are its left and right parts, respectively. First, we make two easy observations on the flags $2 \times 1$ which can be checked manually.
(i) For each $A \in \mathcal{S}$, there exists only one $2 \times 1$ flag $C \in \mathcal{S}$ (possibly $C=A$ ) such that $A$ and $C$ cannot form a $2 \times 2$ square with monochrome diagonal (for part BB, that is YY, and for BY that is YB).
(ii) Let $A_{1}, A_{2}, A_{3} \in \mathcal{S}$ be three distinct elements; then two of them can form a $2 \times 2$ square with yellow diagonal, and two of them can form a $2 \times 2$ square with blue diagonal (for all parts but BB, a pair (BY, YB) fits for both statements, while for all parts but BY, these pairs are (YB, YY) and (BB, YB)).

Now, let $\ell$ and $r$ be the numbers of distinct left and right parts of our 5 flags, respectively. The total number of flags is $5 \leq r \ell$, hence one of the factors (say, $r$ ) should be at least 3. On the other hand, $\ell, r \leq 4$, so there are two flags with coinciding right part; let them be $L_{1} R_{1}$ and $L_{2} R_{1}\left(L_{1} \neq L_{2}\right)$.

Next, since $r \geq 3$, there exist some flags $L_{3} R_{3}$ and $L_{4} R_{4}$ such that $R_{1}, R_{3}, R_{4}$ are distinct. Let $L^{\prime} R^{\prime}$ be the remaining flag. By (i), one of the pairs ( $L^{\prime}, L_{1}$ ) and ( $L^{\prime}, L_{2}$ ) can form a $2 \times 2$ square with monochrome diagonal; we can assume that $L^{\prime}, L_{2}$ form a square with a blue diagonal. Finally, the right parts of two of the flags $L_{1} R_{1}, L_{3} R_{3}, L_{4} R_{4}$ can also form a $2 \times 2$ square with a blue diagonal by (ii). Putting these $2 \times 2$ squares on the diagonal of a $4 \times 4$ square, we find a desired arrangement of four flags.

We are ready to prove the problem statement by induction on $N$; actually, above we have proved the base case $N=4$. For the induction step, assume that $N>4$, consider any $2^{N-2}+1$ flags of length $N$, and arrange them into a large flag of size $\left(2^{N-2}+1\right) \times N$. This flag contains a non-monochrome column since the flags are distinct; we may assume that this column is the first one. By the pigeonhole principle, this column contains at least $\left\lceil\frac{2^{N-2}+1}{2}\right\rceil=2^{N-3}+1$ squares of one color (say, blue). We call the flags with a blue first square good.

Consider all the good flags and remove the first square from each of them. We obtain at least $2^{N-3}+1 \geq M_{N-1}$ flags of length $N-1$; by the induction hypothesis, $N-1$ of them
can form a square $Q$ with the monochrome diagonal. Now, returning the removed squares, we obtain a rectangle $(N-1) \times N$, and our aim is to supplement it on the top by one more flag.

If $Q$ has a yellow diagonal, then we can take each flag with a yellow first square (it exists by a choice of the first column; moreover, it is not used in $Q$ ). Conversely, if the diagonal of $Q$ is blue then we can take any of the $\geq 2^{N-3}+1-(N-1)>0$ remaining good flags. So, in both cases we get a desired $N \times N$ square.

Solution 2. We present a different proof of the estimate $M_{N} \leq 2^{N-2}+1$. We do not use the induction, involving Hall's lemma on matchings instead.

Consider arbitrary $2^{N-2}+1$ distinct flags and arrange them into a large $\left(2^{N-2}+1\right) \times N$ flag. Construct two bipartite graphs $G_{\mathrm{y}}=\left(V \cup V^{\prime}, E_{\mathrm{y}}\right)$ and $G_{\mathrm{b}}=\left(V \cup V^{\prime}, E_{\mathrm{b}}\right)$ with the common set of vertices as follows. Let $V$ and $V^{\prime}$ be the set of columns and the set of flags under consideration, respectively. Next, let the edge $(c, f)$ appear in $E_{y}$ if the intersection of column $c$ and flag $f$ is yellow, and $(c, f) \in E_{\mathrm{b}}$ otherwise. Then we have to prove exactly that one of the graphs $G_{\mathrm{y}}$ and $G_{\mathrm{b}}$ contains a matching with all the vertices of $V$ involved.

Assume that these matchings do not exist. By Hall's lemma, it means that there exist two sets of columns $S_{\mathrm{y}}, S_{\mathrm{b}} \subset V$ such that $\left|E_{\mathrm{y}}\left(S_{\mathrm{y}}\right)\right| \leq\left|S_{\mathrm{y}}\right|-1$ and $\left|E_{\mathrm{b}}\left(S_{\mathrm{b}}\right)\right| \leq\left|S_{\mathrm{b}}\right|-1$ (in the left-hand sides, $E_{\mathrm{y}}\left(S_{\mathrm{y}}\right)$ and $E_{\mathrm{b}}\left(S_{\mathrm{b}}\right)$ denote respectively the sets of all vertices connected to $S_{\mathrm{y}}$ and $S_{\mathrm{b}}$ in the corresponding graphs). Our aim is to prove that this is impossible. Note that $S_{\mathrm{y}}, S_{\mathrm{b}} \neq V$ since $N \leq 2^{N-2}+1$.

First, suppose that $S_{\mathrm{y}} \cap S_{\mathrm{b}} \neq \varnothing$, so there exists some $c \in S_{\mathrm{y}} \cap S_{\mathrm{b}}$. Note that each flag is connected to $c$ either in $G_{\mathrm{y}}$ or in $G_{\mathrm{b}}$, hence $E_{\mathrm{y}}\left(S_{\mathrm{y}}\right) \cup E_{\mathrm{b}}\left(S_{\mathrm{b}}\right)=V^{\prime}$. Hence we have $2^{N-2}+1=\left|V^{\prime}\right| \leq\left|E_{\mathrm{y}}\left(S_{\mathrm{y}}\right)\right|+\left|E_{\mathrm{b}}\left(S_{\mathrm{b}}\right)\right| \leq\left|S_{\mathrm{y}}\right|+\left|S_{\mathrm{b}}\right|-2 \leq 2 N-4$; this is impossible for $N \geq 4$.

So, we have $S_{\mathrm{y}} \cap S_{\mathrm{b}}=\varnothing$. Let $y=\left|S_{\mathrm{y}}\right|, b=\left|S_{\mathrm{b}}\right|$. From the construction of our graph, we have that all the flags in the set $V^{\prime \prime}=V^{\prime} \backslash\left(E_{\mathrm{y}}\left(S_{\mathrm{y}}\right) \cup E_{\mathrm{b}}\left(S_{\mathrm{b}}\right)\right)$ have blue squares in the columns of $S_{\mathrm{y}}$ and yellow squares in the columns of $S_{\mathrm{b}}$. Hence the only undetermined positions in these flags are the remaining $N-y-b$ ones, so $2^{N-y-b} \geq\left|V^{\prime \prime}\right| \geq\left|V^{\prime}\right|-\left|E_{\mathrm{y}}\left(S_{\mathrm{y}}\right)\right|-\left|E_{\mathrm{b}}\left(S_{\mathrm{b}}\right)\right| \geq$ $2^{N-2}+1-(y-1)-(b-1)$, or, denoting $c=y+b, 2^{N-c}+c>2^{N-2}+2$. This is impossible since $N \geq c \geq 2$.

C3. 2500 chess kings have to be placed on a $100 \times 100$ chessboard so that
(i) no king can capture any other one (i.e. no two kings are placed in two squares sharing a common vertex);
(ii) each row and each column contains exactly 25 kings.

Find the number of such arrangements. (Two arrangements differing by rotation or symmetry are supposed to be different.)
(Russia)
Answer. There are two such arrangements.
Solution. Suppose that we have an arrangement satisfying the problem conditions. Divide the board into $2 \times 2$ pieces; we call these pieces blocks. Each block can contain not more than one king (otherwise these two kings would attack each other); hence, by the pigeonhole principle each block must contain exactly one king.

Now assign to each block a letter T or B if a king is placed in its top or bottom half, respectively. Similarly, assign to each block a letter L or R if a king stands in its left or right half. So we define $T$-blocks, $B$-blocks, $L$-blocks, and $R$-blocks. We also combine the letters; we call a block a TL-block if it is simultaneously T-block and L-block. Similarly we define TR-blocks, $B L$-blocks, and BR-blocks. The arrangement of blocks determines uniquely the arrangement of kings; so in the rest of the solution we consider the $50 \times 50$ system of blocks (see Fig. 1). We identify the blocks by their coordinate pairs; the pair $(i, j)$, where $1 \leq i, j \leq 50$, refers to the $j$ th block in the $i$ th row (or the $i$ th block in the $j$ th column). The upper-left block is $(1,1)$.

The system of blocks has the following properties..
( $\mathrm{i}^{\prime}$ ) If $(i, j)$ is a B-block then $(i+1, j)$ is a B-block: otherwise the kings in these two blocks can take each other. Similarly: if $(i, j)$ is a T-block then $(i-1, j)$ is a T-block; if $(i, j)$ is an L-block then $(i, j-1)$ is an L-block; if $(i, j)$ is an R-block then $(i, j+1)$ is an R-block.
(ii') Each column contains exactly 25 L-blocks and 25 R-blocks, and each row contains exactly 25 T-blocks and 25 B-blocks. In particular, the total number of L-blocks (or R-blocks, or T-blocks, or B-blocks) is equal to $25 \cdot 50=1250$.

Consider any B-block of the form $(1, j)$. By ( $\mathrm{i}^{\prime}$ ), all blocks in the $j$ th column are B-blocks; so we call such a column $B$-column. By (ii'), we have 25 B -blocks in the first row, so we obtain 25 B-columns. These 25 B-columns contain 1250 B-blocks, hence all blocks in the remaining columns are T-blocks, and we obtain 25 T-columns. Similarly, there are exactly 25 L-rows and exactly $25 R$-rows.

Now consider an arbitrary pair of a T-column and a neighboring B-column (columns with numbers $j$ and $j+1$ ).


Fig. 1


Fig. 2

Case 1. Suppose that the $j$ th column is a T-column, and the $(j+1)$ th column is a Bcolumn. Consider some index $i$ such that the $i$ th row is an L-row; then $(i, j+1)$ is a BL-block. Therefore, $(i+1, j)$ cannot be a TR-block (see Fig. 2), hence $(i+1, j)$ is a TL-block, thus the
$(i+1)$ th row is an L-row. Now, choosing the $i$ th row to be the topmost L-row, we successively obtain that all rows from the $i$ th to the 50 th are L-rows. Since we have exactly 25 L-rows, it follows that the rows from the 1 st to the 25 th are R-rows, and the rows from the 26 th to the 50th are L-rows.

Now consider the neighboring R-row and L-row (that are the rows with numbers 25 and 26). Replacing in the previous reasoning rows by columns and vice versa, the columns from the 1 st to the 25 th are T-columns, and the columns from the 26 th to the 50 th are B-columns. So we have a unique arrangement of blocks that leads to the arrangement of kings satisfying the condition of the problem (see Fig. 3).


Fig. 3


Fig. 4

Case 2. Suppose that the $j$ th column is a B-column, and the $(j+1)$ th column is a T-column. Repeating the arguments from Case 1, we obtain that the rows from the 1st to the 25th are L-rows (and all other rows are R-rows), the columns from the 1st to the 25 th are B-columns (and all other columns are T-columns), so we find exactly one more arrangement of kings (see Fig. 4).
$\mathbf{C 4}$. Six stacks $S_{1}, \ldots, S_{6}$ of coins are standing in a row. In the beginning every stack contains a single coin. There are two types of allowed moves:
Move 1: If stack $S_{k}$ with $1 \leq k \leq 5$ contains at least one coin, you may remove one coin from $S_{k}$ and add two coins to $S_{k+1}$.
Move 2: If stack $S_{k}$ with $1 \leq k \leq 4$ contains at least one coin, then you may remove one coin from $S_{k}$ and exchange stacks $S_{k+1}$ and $S_{k+2}$.
Decide whether it is possible to achieve by a sequence of such moves that the first five stacks are empty, whereas the sixth stack $S_{6}$ contains exactly $2010^{2010^{2010}}$ coins.
$\mathbf{C 4}{ }^{\prime}$. Same as Problem C4, but the constant $2010^{2010^{2010}}$ is replaced by $2010^{2010}$.
(Netherlands)
Answer. Yes (in both variants of the problem). There exists such a sequence of moves.
Solution. Denote by $\left(a_{1}, a_{2}, \ldots, a_{n}\right) \rightarrow\left(a_{1}^{\prime}, a_{2}^{\prime}, \ldots, a_{n}^{\prime}\right)$ the following: if some consecutive stacks contain $a_{1}, \ldots, a_{n}$ coins, then it is possible to perform several allowed moves such that the stacks contain $a_{1}^{\prime}, \ldots, a_{n}^{\prime}$ coins respectively, whereas the contents of the other stacks remain unchanged.

Let $A=2010^{2010}$ or $A=2010^{2010^{2010}}$, respectively. Our goal is to show that

$$
(1,1,1,1,1,1) \rightarrow(0,0,0,0,0, A)
$$

First we prove two auxiliary observations.
Lemma 1. $(a, 0,0) \rightarrow\left(0,2^{a}, 0\right)$ for every $a \geq 1$.
Proof. We prove by induction that $(a, 0,0) \rightarrow\left(a-k, 2^{k}, 0\right)$ for every $1 \leq k \leq a$. For $k=1$, apply Move 1 to the first stack:

$$
(a, 0,0) \rightarrow(a-1,2,0)=\left(a-1,2^{1}, 0\right)
$$

Now assume that $k<a$ and the statement holds for some $k<a$. Starting from $\left(a-k, 2^{k}, 0\right)$, apply Move 1 to the middle stack $2^{k}$ times, until it becomes empty. Then apply Move 2 to the first stack:

$$
\left(a-k, 2^{k}, 0\right) \rightarrow\left(a-k, 2^{k}-1,2\right) \rightarrow \cdots \rightarrow\left(a-k, 0,2^{k+1}\right) \rightarrow\left(a-k-1,2^{k+1}, 0\right)
$$

Hence,

$$
(a, 0,0) \rightarrow\left(a-k, 2^{k}, 0\right) \rightarrow\left(a-k-1,2^{k+1}, 0\right)
$$

Lemma 2. For every positive integer $n$, let $P_{n}=\underbrace{2^{2 \cdot b^{2}}}_{n}$ (e.g. $P_{3}=2^{2^{2}}=16$ ). Then $(a, 0,0,0) \rightarrow\left(0, P_{a}, 0,0\right)$ for every $a \geq 1$.
Proof. Similarly to Lemma 1 , we prove that $(a, 0,0,0) \rightarrow\left(a-k, P_{k}, 0,0\right)$ for every $1 \leq k \leq a$.
For $k=1$, apply Move 1 to the first stack:

$$
(a, 0,0,0) \rightarrow(a-1,2,0,0)=\left(a-1, P_{1}, 0,0\right)
$$

Now assume that the lemma holds for some $k<a$. Starting from ( $a-k, P_{k}, 0,0$ ), apply Lemma 1, then apply Move 1 to the first stack:

$$
\left(a-k, P_{k}, 0,0\right) \rightarrow\left(a-k, 0,2^{P_{k}}, 0\right)=\left(a-k, 0, P_{k+1}, 0\right) \rightarrow\left(a-k-1, P_{k+1}, 0,0\right)
$$

Therefore,

$$
(a, 0,0,0) \rightarrow\left(a-k, P_{k}, 0,0\right) \rightarrow\left(a-k-1, P_{k+1}, 0,0\right)
$$

Now we prove the statement of the problem.
First apply Move 1 to stack 5 , then apply Move 2 to stacks $S_{4}, S_{3}, S_{2}$ and $S_{1}$ in this order. Then apply Lemma 2 twice:

$$
\begin{gathered}
(1,1,1,1,1,1) \rightarrow(1,1,1,1,0,3) \rightarrow(1,1,1,0,3,0) \rightarrow(1,1,0,3,0,0) \rightarrow(1,0,3,0,0,0) \rightarrow \\
\quad \rightarrow(0,3,0,0,0,0) \rightarrow\left(0,0, P_{3}, 0,0,0\right)=(0,0,16,0,0,0) \rightarrow\left(0,0,0, P_{16}, 0,0\right) .
\end{gathered}
$$

We already have more than $A$ coins in stack $S_{4}$, since

$$
A \leq 2010^{2010^{2010}}<\left(2^{11}\right)^{2010^{2010}}=2^{11 \cdot 2010^{2010}}<2^{20100^{2011}}<2^{\left(2^{11}\right)^{2011}}=2^{2^{11 \cdot 2011}}<2^{2^{2^{15}}}<P_{16}
$$

To decrease the number of coins in stack $S_{4}$, apply Move 2 to this stack repeatedly until its size decreases to $A / 4$. (In every step, we remove a coin from $S_{4}$ and exchange the empty stacks $S_{5}$ and $S_{6}$.)

$$
\begin{aligned}
\left(0,0,0, P_{16}, 0,0\right) \rightarrow & \left(0,0,0, P_{16}-1,0,0\right) \rightarrow\left(0,0,0, P_{16}-2,0,0\right) \rightarrow \\
& \rightarrow \cdots \rightarrow(0,0,0, A / 4,0,0) .
\end{aligned}
$$

Finally, apply Move 1 repeatedly to empty stacks $S_{4}$ and $S_{5}$ :

$$
(0,0,0, A / 4,0,0) \rightarrow \cdots \rightarrow(0,0,0,0, A / 2,0) \rightarrow \cdots \rightarrow(0,0,0,0,0, A)
$$

Comment 1. Starting with only 4 stack, it is not hard to check manually that we can achieve at most 28 coins in the last position. However, around 5 and 6 stacks the maximal number of coins explodes. With 5 stacks it is possible to achieve more than $2^{2^{14}}$ coins. With 6 stacks the maximum is greater than $P_{P_{2^{14}}}$.

It is not hard to show that the numbers $2010^{2010}$ and $2010^{2010^{2010}}$ in the problem can be replaced by any nonnegative integer up to $P_{P_{2} 14}$.
Comment 2. The simpler variant $\mathrm{C} 4^{\prime}$ of the problem can be solved without Lemma 2:

$$
\begin{aligned}
(1,1,1,1,1,1) & \rightarrow(0,3,1,1,1,1) \rightarrow(0,1,5,1,1,1) \rightarrow(0,1,1,9,1,1) \rightarrow \\
& \rightarrow(0,1,1,1,17,1) \rightarrow(0,1,1,1,0,35) \rightarrow(0,1,1,0,35,0) \rightarrow(0,1,0,35,0,0) \rightarrow \\
& \rightarrow(0,0,35,0,0,0) \rightarrow\left(0,0,1,2^{34}, 0,0\right) \rightarrow\left(0,0,1,0,2^{2^{34}}, 0\right) \rightarrow\left(0,0,0,2^{2^{34}}, 0,0\right) \\
& \rightarrow\left(0,0,0,2^{2^{34}}-1,0,0\right) \rightarrow \ldots \rightarrow(0,0,0, A / 4,0,0) \rightarrow(0,0,0,0, A / 2,0) \rightarrow(0,0,0,0,0, A) .
\end{aligned}
$$

For this reason, the PSC suggests to consider the problem C4 as well. Problem C4 requires more invention and technical care. On the other hand, the problem statement in C 4 ' hides the fact that the resulting amount of coins can be such incredibly huge and leaves this discovery to the students.

C5. $n \geq 4$ players participated in a tennis tournament. Any two players have played exactly one game, and there was no tie game. We call a company of four players bad if one player was defeated by the other three players, and each of these three players won a game and lost another game among themselves. Suppose that there is no bad company in this tournament. Let $w_{i}$ and $\ell_{i}$ be respectively the number of wins and losses of the $i$ th player. Prove that

$$
\begin{equation*}
\sum_{i=1}^{n}\left(w_{i}-\ell_{i}\right)^{3} \geq 0 \tag{1}
\end{equation*}
$$

(South Korea)
Solution. For any tournament $T$ satisfying the problem condition, denote by $S(T)$ sum under consideration, namely

$$
S(T)=\sum_{i=1}^{n}\left(w_{i}-\ell_{i}\right)^{3}
$$

First, we show that the statement holds if a tournament $T$ has only 4 players. Actually, let $A=\left(a_{1}, a_{2}, a_{3}, a_{4}\right)$ be the number of wins of the players; we may assume that $a_{1} \geq a_{2} \geq a_{3} \geq a_{4}$. We have $a_{1}+a_{2}+a_{3}+a_{4}=\binom{4}{2}=6$, hence $a_{4} \leq 1$. If $a_{4}=0$, then we cannot have $a_{1}=a_{2}=a_{3}=2$, otherwise the company of all players is bad. Hence we should have $A=(3,2,1,0)$, and $S(T)=3^{3}+1^{3}+(-1)^{3}+(-3)^{3}=0$. On the other hand, if $a_{4}=1$, then only two possibilities, $A=(3,1,1,1)$ and $A=(2,2,1,1)$ can take place. In the former case we have $S(T)=3^{3}+3 \cdot(-2)^{3}>0$, while in the latter one $S(T)=1^{3}+1^{3}+(-1)^{3}+(-1)^{3}=0$, as desired.

Now we turn to the general problem. Consider a tournament $T$ with no bad companies and enumerate the players by the numbers from 1 to $n$. For every 4 players $i_{1}, i_{2}, i_{3}, i_{4}$ consider a "sub-tournament" $T_{i_{1} i_{2} i_{3} i_{4}}$ consisting of only these players and the games which they performed with each other. By the abovementioned, we have $S\left(T_{i_{1} i_{2} i_{3} i_{4}}\right) \geq 0$. Our aim is to prove that

$$
\begin{equation*}
S(T)=\sum_{i_{1}, i_{2}, i_{3}, i_{4}} S\left(T_{i_{1} i_{2} i_{3} i_{4}}\right) \tag{2}
\end{equation*}
$$

where the sum is taken over all 4 -tuples of distinct numbers from the set $\{1, \ldots, n\}$. This way the problem statement will be established.

We interpret the number $\left(w_{i}-\ell_{i}\right)^{3}$ as following. For $i \neq j$, let $\varepsilon_{i j}=1$ if the $i$ th player wins against the $j$ th one, and $\varepsilon_{i j}=-1$ otherwise. Then

$$
\left(w_{i}-\ell_{i}\right)^{3}=\left(\sum_{j \neq i} \varepsilon_{i j}\right)^{3}=\sum_{j_{1}, j_{2}, j_{3} \neq i} \varepsilon_{i j_{1}} \varepsilon_{i j_{2}} \varepsilon_{i j_{3}}
$$

Hence,

$$
S(T)=\sum_{i \notin\left\{j_{1}, j_{2}, j_{3}\right\}} \varepsilon_{i j_{1}} \varepsilon_{i j_{2}} \varepsilon_{i j_{3}} .
$$

To simplify this expression, consider all the terms in this sum where two indices are equal. If, for instance, $j_{1}=j_{2}$, then the term contains $\varepsilon_{i j_{1}}^{2}=1$, so we can replace this term by $\varepsilon_{i j_{3}}$. Make such replacements for each such term; obviously, after this change each term of the form $\varepsilon_{i j_{3}}$ will appear $P(T)$ times, hence

$$
S(T)=\sum_{\left|\left\{i, j_{1}, j_{2}, j_{3}\right\}\right|=4} \varepsilon_{i j_{1}} \varepsilon_{i j_{2}} \varepsilon_{i j_{3}}+P(T) \sum_{i \neq j} \varepsilon_{i j}=S_{1}(T)+P(T) S_{2}(T)
$$

We show that $S_{2}(T)=0$ and hence $S(T)=S_{1}(T)$ for each tournament. Actually, note that $\varepsilon_{i j}=-\varepsilon_{j i}$, and the whole sum can be split into such pairs. Since the sum in each pair is 0 , so is $S_{2}(T)$.

Thus the desired equality (2) rewrites as

$$
\begin{equation*}
S_{1}(T)=\sum_{i_{1}, i_{2}, i_{3}, i_{4}} S_{1}\left(T_{i_{1} i_{2} i_{3} i_{4}}\right) \tag{3}
\end{equation*}
$$

Now, if all the numbers $j_{1}, j_{2}, j_{3}$ are distinct, then the set $\left\{i, j_{1}, j_{2}, j_{3}\right\}$ is contained in exactly one 4 -tuple, hence the term $\varepsilon_{i j_{1}} \varepsilon_{i j_{2}} \varepsilon_{i j_{3}}$ appears in the right-hand part of (3) exactly once, as well as in the left-hand part. Clearly, there are no other terms in both parts, so the equality is established.

Solution 2. Similarly to the first solution, we call the subsets of players as companies, and the $k$-element subsets will be called as $k$-companies.

In any company of the players, call a player the local champion of the company if he defeated all other members of the company. Similarly, if a player lost all his games against the others in the company then call him the local loser of the company. Obviously every company has at most one local champion and at most one local loser. By the condition of the problem, whenever a 4-company has a local loser, then this company has a local champion as well.

Suppose that $k$ is some positive integer, and let us count all cases when a player is the local champion of some $k$-company. The $i$ th player won against $w_{i}$ other player. To be the local champion of a $k$-company, he must be a member of the company, and the other $k-1$ members must be chosen from those whom he defeated. Therefore, the $i$ th player is the local champion of $\binom{w_{i}}{k-1} k$-companies. Hence, the total number of local champions of all $k$-companies is $\sum_{i=1}^{n}\binom{w_{i}}{k-1}$.

Similarly, the total number of local losers of the $k$-companies is $\sum_{i=1}^{n}\binom{\ell_{i}}{k-1}$.
Now apply this for $k=2,3$ and 4 .
Since every game has a winner and a loser, we have $\sum_{i=1}^{n} w_{i}=\sum_{i=1}^{n} \ell_{i}=\binom{n}{2}$, and hence

$$
\begin{equation*}
\sum_{i=1}^{n}\left(w_{i}-\ell_{i}\right)=0 \tag{4}
\end{equation*}
$$

In every 3-company, either the players defeated one another in a cycle or the company has both a local champion and a local loser. Therefore, the total number of local champions and local losers in the 3-companies is the same, $\sum_{i=1}^{n}\binom{w_{i}}{2}=\sum_{i=1}^{n}\binom{\ell_{i}}{2}$. So we have

$$
\begin{equation*}
\sum_{i=1}^{n}\left(\binom{w_{i}}{2}-\binom{\ell_{i}}{2}\right)=0 \tag{5}
\end{equation*}
$$

In every 4-company, by the problem's condition, the number of local losers is less than or equal to the number of local champions. Then the same holds for the total numbers of local
champions and local losers in all 4-companies, so $\sum_{i=1}^{n}\binom{w_{i}}{3} \geq \sum_{i=1}^{n}\binom{\ell_{i}}{3}$. Hence,

$$
\begin{equation*}
\sum_{i=1}^{n}\left(\binom{w_{i}}{3}-\binom{\ell_{i}}{3}\right) \geq 0 . \tag{6}
\end{equation*}
$$

Now we establish the problem statement (1) as a linear combination of (4), (5) and (6). It is easy check that

$$
(x-y)^{3}=24\left(\binom{x}{3}-\binom{y}{3}\right)+24\left(\binom{x}{2}-\binom{y}{2}\right)-\left(3(x+y)^{2}-4\right)(x-y) .
$$

Apply this identity to $x=w_{1}$ and $y=\ell_{i}$. Since every player played $n-1$ games, we have $w_{i}+\ell_{i}=n-1$, and thus

$$
\left(w_{i}-\ell_{i}\right)^{3}=24\left(\binom{w_{i}}{3}-\binom{\ell_{i}}{3}\right)+24\left(\binom{w_{i}}{2}-\binom{\ell_{i}}{2}\right)-\left(3(n-1)^{2}-4\right)\left(w_{i}-\ell_{i}\right) .
$$

Then

$$
\sum_{i=1}^{n}\left(w_{i}-\ell_{i}\right)^{3}=24 \underbrace{\sum_{i=1}^{n}\left(\binom{w_{i}}{3}-\binom{\ell_{i}}{3}\right)}_{\geq 0}+24 \underbrace{\sum_{i=1}^{n}\left(\binom{w_{i}}{2}-\binom{\ell_{i}}{2}\right)}_{0}-\left(3(n-1)^{2}-4\right) \underbrace{\sum_{i=1}^{n}\left(w_{i}-\ell_{i}\right)}_{0} \geq 0
$$

C6. Given a positive integer $k$ and other two integers $b>w>1$. There are two strings of pearls, a string of $b$ black pearls and a string of $w$ white pearls. The length of a string is the number of pearls on it.

One cuts these strings in some steps by the following rules. In each step:
(i) The strings are ordered by their lengths in a non-increasing order. If there are some strings of equal lengths, then the white ones precede the black ones. Then $k$ first ones (if they consist of more than one pearl) are chosen; if there are less than $k$ strings longer than 1 , then one chooses all of them.
(ii) Next, one cuts each chosen string into two parts differing in length by at most one.
(For instance, if there are strings of $5,4,4,2$ black pearls, strings of $8,4,3$ white pearls and $k=4$, then the strings of 8 white, 5 black, 4 white and 4 black pearls are cut into the parts $(4,4),(3,2),(2,2)$ and $(2,2)$, respectively.)

The process stops immediately after the step when a first isolated white pearl appears. Prove that at this stage, there will still exist a string of at least two black pearls.
(Canada)
Solution 1. Denote the situation after the $i$ th step by $A_{i}$; hence $A_{0}$ is the initial situation, and $A_{i-1} \rightarrow A_{i}$ is the $i$ th step. We call a string containing $m$ pearls an $m$-string; it is an $m$-w-string or a $m$-b-string if it is white or black, respectively.

We continue the process until every string consists of a single pearl. We will focus on three moments of the process: (a) the first stage $A_{s}$ when the first 1 -string (no matter black or white) appears; (b) the first stage $A_{t}$ where the total number of strings is greater than $k$ (if such moment does not appear then we put $t=\infty$ ); and (c) the first stage $A_{f}$ when all black pearls are isolated. It is sufficient to prove that in $A_{f-1}$ (or earlier), a 1-w-string appears.

We start with some easy properties of the situations under consideration. Obviously, we have $s \leq f$. Moreover, all b-strings from $A_{f-1}$ become single pearls in the $f$ th step, hence all of them are 1 - or 2 -b-strings.

Next, observe that in each step $A_{i} \rightarrow A_{i+1}$ with $i \leq t-1$, all $(>1)$-strings were cut since there are not more than $k$ strings at all; if, in addition, $i<s$, then there were no 1 -string, so all the strings were cut in this step.

Now, let $B_{i}$ and $b_{i}$ be the lengths of the longest and the shortest b-strings in $A_{i}$, and let $W_{i}$ and $w_{i}$ be the same for w-strings. We show by induction on $i \leq \min \{s, t\}$ that (i) the situation $A_{i}$ contains exactly $2^{i}$ black and $2^{i}$ white strings, (ii) $B_{i} \geq W_{i}$, and (iii) $b_{i} \geq w_{i}$. The base case $i=0$ is obvious. For the induction step, if $i \leq \min \{s, t\}$ then in the $i$ th step, each string is cut, thus the claim (i) follows from the induction hypothesis; next, we have $B_{i}=\left\lceil B_{i-1} / 2\right\rceil \geq\left\lceil W_{i-1} / 2\right\rceil=W_{i}$ and $b_{i}=\left\lfloor b_{i-1} / 2\right\rfloor \geq\left\lfloor w_{i-1} / 2\right\rfloor=w_{i}$, thus establishing (ii) and (iii).

For the numbers $s, t, f$, two cases are possible.
Case 1. Suppose that $s \leq t$ or $f \leq t+1$ (and hence $s \leq t+1$ ); in particular, this is true when $t=\infty$. Then in $A_{s-1}$ we have $B_{s-1} \geq W_{s-1}, b_{s-1} \geq w_{s-1}>1$ as $s-1 \leq \min \{s, t\}$. Now, if $s=f$, then in $A_{s-1}$, there is no 1 -w-string as well as no ( $>2$ )-b-string. That is, $2=B_{s-1} \geq W_{s-1} \geq b_{s-1} \geq w_{s-1}>1$, hence all these numbers equal 2. This means that in $A_{s-1}$, all strings contain 2 pearls, and there are $2^{s-1}$ black and $2^{s-1}$ white strings, which means $b=2 \cdot 2^{s-1}=w$. This contradicts the problem conditions.

Hence we have $s \leq f-1$ and thus $s \leq t$. Therefore, in the $s$ th step each string is cut into two parts. Now, if a 1-b-string appears in this step, then from $w_{s-1} \leq b_{s-1}$ we see that a

1 -w-string appears as well; so, in each case in the sth step a 1 -w-string appears, while not all black pearls become single, as desired.

Case 2. Now assume that $t+1 \leq s$ and $t+2 \leq f$. Then in $A_{t}$ we have exactly $2^{t}$ white and $2^{t}$ black strings, all being larger than 1 , and $2^{t+1}>k \geq 2^{t}$ (the latter holds since $2^{t}$ is the total number of strings in $A_{t-1}$ ). Now, in the $(t+1)$ st step, exactly $k$ strings are cut, not more than $2^{t}$ of them being black; so the number of w-strings in $A_{t+1}$ is at least $2^{t}+\left(k-2^{t}\right)=k$. Since the number of w-strings does not decrease in our process, in $A_{f-1}$ we have at least $k$ white strings as well.

Finally, in $A_{f-1}$, all b-strings are not larger than 2, and at least one 2-b-string is cut in the $f$ th step. Therefore, at most $k-1$ white strings are cut in this step, hence there exists a w-string $\mathcal{W}$ which is not cut in the $f$ th step. On the other hand, since a 2 -b-string is cut, all $(\geq 2)$-w-strings should also be cut in the $f$ th step; hence $\mathcal{W}$ should be a single pearl. This is exactly what we needed.
Comment. In this solution, we used the condition $b \neq w$ only to avoid the case $b=w=2^{t}$. Hence, if a number $b=w$ is not a power of 2 , then the problem statement is also valid.

Solution 2. We use the same notations as introduced in the first paragraph of the previous solution. We claim that at every stage, there exist a $u$-b-string and a $v$-w-string such that either
(i) $u>v \geq 1$, or
(ii) $2 \leq u \leq v<2 u$, and there also exist $k-1$ of ( $>v / 2$ )-strings other than considered above.

First, we notice that this statement implies the problem statement. Actually, in both cases (i) and (ii) we have $u>1$, so at each stage there exists a ( $\geq 2$ )-b-string, and for the last stage it is exactly what we need.

Now, we prove the claim by induction on the number of the stage. Obviously, for $A_{0}$ the condition (i) holds since $b>w$. Further, we suppose that the statement holds for $A_{i}$, and prove it for $A_{i+1}$. Two cases are possible.

Case 1. Assume that in $A_{i}$, there are a $u$-b-string and a $v$-w-string with $u>v$. We can assume that $v$ is the length of the shortest w-string in $A_{i}$; since we are not at the final stage, we have $v \geq 2$. Now, in the $(i+1)$ st step, two subcases may occur.

Subcase 1a. Suppose that either no $u$-b-string is cut, or both some $u$-b-string and some $v$-w-string are cut. Then in $A_{i+1}$, we have either a $u$-b-string and a $(\leq v)$-w-string (and (i) is valid), or we have a [u/2]-b-string and a $\lfloor v / 2\rfloor$-w-string. In the latter case, from $u>v$ we get $\lceil u / 2\rceil>\lfloor v / 2\rfloor$, and (i) is valid again.

Subcase 1 . Now, some $u$-b-string is cut, and no $v$-w-string is cut (and hence all the strings which are cut are longer than $v$ ). If $u^{\prime}=\lceil u / 2\rceil>v$, then the condition (i) is satisfied since we have a $u^{\prime}$-b-string and a $v$-w-string in $A_{i+1}$. Otherwise, notice that the inequality $u>v \geq 2$ implies $u^{\prime} \geq 2$. Furthermore, besides a fixed $u$-b-string, other $k-1$ of $(\geq v+1)$-strings should be cut in the $(i+1)$ st step, hence providing at least $k-1$ of $(\geq\lceil(v+1) / 2\rceil)$-strings, and $\lceil(v+1) / 2\rceil>v / 2$. So, we can put $v^{\prime}=v$, and we have $u^{\prime} \leq v<u \leq 2 u^{\prime}$, so the condition (ii) holds for $A_{i+1}$.

Case 2. Conversely, assume that in $A_{i}$ there exist a $u$-b-string, a $v$-w-string $(2 \leq u \leq v<2 u)$ and a set $S$ of $k-1$ other strings larger than $v / 2$ (and hence larger than 1 ). In the ( $i+1$ )st step, three subcases may occur.

Subcase 2a. Suppose that some $u$-b-string is not cut, and some $v$-w-string is cut. The latter one results in a $\lfloor v / 2\rfloor$-w-string, we have $v^{\prime}=\lfloor v / 2\rfloor<u$, and the condition (i) is valid.

Subcase 2b. Next, suppose that no $v$-w-string is cut (and therefore no $u$-b-string is cut as $u \leq v)$. Then all $k$ strings which are cut have the length $>v$, so each one results in a ( $>v / 2$ )string. Hence in $A_{i+1}$, there exist $k \geq k-1$ of $(>v / 2)$-strings other than the considered $u$ - and $v$-strings, and the condition (ii) is satisfied.

Subcase 2c. In the remaining case, all $u$-b-strings are cut. This means that all $(\geq u)$-strings are cut as well, hence our $v$-w-string is cut. Therefore in $A_{i+1}$ there exists a $\lceil u / 2\rceil$-b-string together with a $\lfloor v / 2\rfloor$-w-string. Now, if $u^{\prime}=\lceil u / 2\rceil>\lfloor v / 2\rfloor=v^{\prime}$ then the condition (i) is fulfilled. Otherwise, we have $u^{\prime} \leq v^{\prime}<u \leq 2 u^{\prime}$. In this case, we show that $u^{\prime} \geq 2$. If, to the contrary, $u^{\prime}=1$ (and hence $u=2$ ), then all black and white ( $\geq 2$ )-strings should be cut in the $(i+1)$ st step, and among these strings there are at least a $u$-b-string, a $v$-w-string, and $k-1$ strings in $S(k+1$ strings altogether). This is impossible.

Hence, we get $2 \leq u^{\prime} \leq v^{\prime}<2 u^{\prime}$. To reach (ii), it remains to check that in $A_{i+1}$, there exists a set $S^{\prime}$ of $k-1$ other strings larger than $v^{\prime} / 2$. These will be exactly the strings obtained from the elements of $S$. Namely, each $s \in S$ was either cut in the $(i+1)$ st step, or not. In the former case, let us include into $S^{\prime}$ the largest of the strings obtained from $s$; otherwise we include $s$ itself into $S^{\prime}$. All $k-1$ strings in $S^{\prime}$ are greater than $v / 2 \geq v^{\prime}$, as desired.

C7. Let $P_{1}, \ldots, P_{s}$ be arithmetic progressions of integers, the following conditions being satisfied:
(i) each integer belongs to at least one of them;
(ii) each progression contains a number which does not belong to other progressions.

Denote by $n$ the least common multiple of steps of these progressions; let $n=p_{1}^{\alpha_{1}} \ldots p_{k}^{\alpha_{k}}$ be its prime factorization. Prove that

$$
s \geq 1+\sum_{i=1}^{k} \alpha_{i}\left(p_{i}-1\right)
$$

(Germany)
Solution 1. First, we prove the key lemma, and then we show how to apply it to finish the solution.

Let $n_{1}, \ldots, n_{k}$ be positive integers. By an $n_{1} \times n_{2} \times \cdots \times n_{k}$ grid we mean the set $N=$ $\left\{\left(a_{1}, \ldots, a_{k}\right): a_{i} \in \mathbb{Z}, 0 \leq a_{i} \leq n_{i}-1\right\}$; the elements of $N$ will be referred to as points. In this grid, we define $a$ subgrid as a subset of the form

$$
\begin{equation*}
L=\left\{\left(b_{1}, \ldots, b_{k}\right) \in N: b_{i_{1}}=x_{i_{1}}, \ldots, b_{i_{t}}=x_{i_{t}}\right\} \tag{1}
\end{equation*}
$$

where $I=\left\{i_{1}, \ldots, i_{t}\right\}$ is an arbitrary nonempty set of indices, and $x_{i_{j}} \in\left[0, n_{i_{j}}-1\right](1 \leq j \leq t)$ are fixed integer numbers. Further, we say that a subgrid (1) is orthogonal to the $i$ th coordinate axis if $i \in I$, and that it is parallel to the $i$ th coordinate axis otherwise.
Lemma. Assume that the grid $N$ is covered by subgrids $L_{1}, L_{2}, \ldots, L_{s}$ (this means $\left.N=\bigcup_{i=1}^{s} L_{i}\right)$ so that
(ii') each subgrid contains a point which is not covered by other subgrids;
(iii) for each coordinate axis, there exists a subgrid $L_{i}$ orthogonal to this axis.

Then

$$
s \geq 1+\sum_{i=1}^{k}\left(n_{i}-1\right)
$$

Proof. Assume to the contrary that $s \leq \sum_{i}\left(n_{i}-1\right)=s^{\prime}$. Our aim is to find a point that is not covered by $L_{1}, \ldots, L_{s}$.

The idea of the proof is the following. Imagine that we expand each subgrid to some maximal subgrid so that for the $i$ th axis, there will be at most $n_{i}-1$ maximal subgrids orthogonal to this axis. Then the desired point can be found easily: its $i$ th coordinate should be that not covered by the maximal subgrids orthogonal to the $i$ th axis. Surely, the conditions for existence of such expansion are provided by Hall's lemma on matchings. So, we will follow this direction, although we will apply Hall's lemma to some subgraph instead of the whole graph.

Construct a bipartite graph $G=\left(V \cup V^{\prime}, E\right)$ as follows. Let $V=\left\{L_{1}, \ldots, L_{s}\right\}$, and let $V^{\prime}=\left\{v_{i j}: 1 \leq i \leq s, 1 \leq j \leq n_{i}-1\right\}$ be some set of $s^{\prime}$ elements. Further, let the edge ( $L_{m}, v_{i j}$ ) appear iff $L_{m}$ is orthogonal to the $i$ th axis.

For each subset $W \subset V$, denote

$$
f(W)=\left\{v \in V^{\prime}:(L, v) \in E \text { for some } L \in W\right\}
$$

Notice that $f(V)=V^{\prime}$ by (iii).
Now, consider the set $W \subset V$ containing the maximal number of elements such that $|W|>$ $|f(W)|$; if there is no such set then we set $W=\varnothing$. Denote $W^{\prime}=f(W), U=V \backslash W, U^{\prime}=V^{\prime} \backslash W^{\prime}$.

By our assumption and the Lemma condition, $|f(V)|=\left|V^{\prime}\right| \geq|V|$, hence $W \neq V$ and $U \neq \varnothing$. Permuting the coordinates, we can assume that $U^{\prime}=\left\{v_{i j}: 1 \leq i \leq \ell\right\}, W^{\prime}=\left\{v_{i j}: \ell+1 \leq i \leq k\right\}$.

Consider the induced subgraph $G^{\prime}$ of $G$ on the vertices $U \cup U^{\prime}$. We claim that for every $X \subset U$, we get $\left|f(X) \cap U^{\prime}\right| \geq|X|$ (so $G^{\prime}$ satisfies the conditions of Hall's lemma). Actually, we have $|W| \geq|f(W)|$, so if $|X|>\left|f(X) \cap U^{\prime}\right|$ for some $X \subset U$, then we have

$$
|W \cup X|=|W|+|X|>|f(W)|+\left|f(X) \cap U^{\prime}\right|=\left|f(W) \cup\left(f(X) \cap U^{\prime}\right)\right|=|f(W \cup X)|
$$

This contradicts the maximality of $|W|$.
Thus, applying Hall's lemma, we can assign to each $L \in U$ some vertex $v_{i j} \in U^{\prime}$ so that to distinct elements of $U$, distinct vertices of $U^{\prime}$ are assigned. In this situation, we say that $L \in U$ corresponds to the $i$ th axis, and write $g(L)=i$. Since there are $n_{i}-1$ vertices of the form $v_{i j}$, we get that for each $1 \leq i \leq \ell$, not more than $n_{i}-1$ subgrids correspond to the $i$ th axis.

Finally, we are ready to present the desired point. Since $W \neq V$, there exists a point $b=\left(b_{1}, b_{2}, \ldots, b_{k}\right) \in N \backslash\left(\cup_{L \in W} L\right)$. On the other hand, for every $1 \leq i \leq \ell$, consider any subgrid $L \in U$ with $g(L)=i$. This means exactly that $L$ is orthogonal to the $i$ th axis, and hence all its elements have the same $i$ th coordinate $c_{L}$. Since there are at most $n_{i}-1$ such subgrids, there exists a number $0 \leq a_{i} \leq n_{i}-1$ which is not contained in a set $\left\{c_{L}: g(L)=i\right\}$. Choose such number for every $1 \leq i \leq \ell$. Now we claim that point $a=\left(a_{1}, \ldots, a_{\ell}, b_{\ell+1}, \ldots, b_{k}\right)$ is not covered, hence contradicting the Lemma condition.

Surely, point $a$ cannot lie in some $L \in U$, since all the points in $L$ have $g(L)$ th coordinate $c_{L} \neq a_{g(L)}$. On the other hand, suppose that $a \in L$ for some $L \in W$; recall that $b \notin L$. But the points $a$ and $b$ differ only at first $\ell$ coordinates, so $L$ should be orthogonal to at least one of the first $\ell$ axes, and hence our graph contains some edge $\left(L, v_{i j}\right)$ for $i \leq \ell$. It contradicts the definition of $W^{\prime}$. The Lemma is proved.

Now we turn to the problem. Let $d_{j}$ be the step of the progression $P_{j}$. Note that since $n=$ l.c.m. $\left(d_{1}, \ldots, d_{s}\right)$, for each $1 \leq i \leq k$ there exists an index $j(i)$ such that $p_{i}^{\alpha_{i}} \mid d_{j(i)}$. We assume that $n>1$; otherwise the problem statement is trivial.

For each $0 \leq m \leq n-1$ and $1 \leq i \leq k$, let $m_{i}$ be the residue of $m$ modulo $p_{i}^{\alpha_{i}}$, and let $m_{i}=\overline{r_{i \alpha_{i}} \ldots r_{i 1}}$ be the base $p_{i}$ representation of $m_{i}$ (possibly, with some leading zeroes). Now, we put into correspondence to $m$ the sequence $r(m)=\left(r_{11}, \ldots, r_{1 \alpha_{1}}, r_{21}, \ldots, r_{k \alpha_{k}}\right)$. Hence $r(m)$ lies in a $\underbrace{p_{1} \times \cdots \times p_{1}}_{\alpha_{1} \text { times }} \times \cdots \times \underbrace{p_{k} \times \cdots \times p_{k}}_{\alpha_{k} \text { times }}$ grid $N$.

Surely, if $r(m)=r\left(m^{\prime}\right)$ then $p_{i}^{\alpha_{i}} \mid m_{i}-m_{i}^{\prime}$, which follows $p_{i}^{\alpha_{i}} \mid m-m^{\prime}$ for all $1 \leq i \leq k$; consequently, $n \mid m-m^{\prime}$. So, when $m$ runs over the set $\{0, \ldots, n-1\}$, the sequences $r(m)$ do not repeat; since $|N|=n$, this means that $r$ is a bijection between $\{0, \ldots, n-1\}$ and $N$. Now we will show that for each $1 \leq i \leq s$, the set $L_{i}=\left\{r(m): m \in P_{i}\right\}$ is a subgrid, and that for each axis there exists a subgrid orthogonal to this axis. Obviously, these subgrids cover $N$, and the condition (ii') follows directly from (ii). Hence the Lemma provides exactly the estimate we need.

Consider some $1 \leq j \leq s$ and let $d_{j}=p_{1}^{\gamma_{1}} \ldots p_{k}^{\gamma_{k}}$. Consider some $q \in P_{j}$ and let $r(q)=$ $\left(r_{11}, \ldots, r_{k \alpha_{k}}\right)$. Then for an arbitrary $q^{\prime}$, setting $r\left(q^{\prime}\right)=\left(r_{11}^{\prime}, \ldots, r_{k \alpha_{k}}^{\prime}\right)$ we have

$$
q^{\prime} \in P_{j} \quad \Longleftrightarrow p_{i}^{\gamma_{i}} \mid q-q^{\prime} \text { for each } 1 \leq i \leq k \quad \Longleftrightarrow \quad r_{i, t}=r_{i, t}^{\prime} \text { for all } t \leq \gamma_{i}
$$

Hence $L_{j}=\left\{\left(r_{11}^{\prime}, \ldots, r_{k \alpha_{k}}^{\prime}\right) \in N: r_{i, t}=r_{i, t}^{\prime}\right.$ for all $\left.t \leq \gamma_{i}\right\}$ which means that $L_{j}$ is a subgrid containing $r(q)$. Moreover, in $L_{j(i)}$, all the coordinates corresponding to $p_{i}$ are fixed, so it is orthogonal to all of their axes, as desired.

Comment 1. The estimate in the problem is sharp for every $n$. One of the possible examples is the following one. For each $1 \leq i \leq k, 0 \leq j \leq \alpha_{i}-1,1 \leq k \leq p-1$, let

$$
P_{i, j, k}=k p_{i}^{j}+p_{i}^{j+1} \mathbb{Z},
$$

and add the progression $P_{0}=n \mathbb{Z}$. One can easily check that this set satisfies all the problem conditions. There also exist other examples.

On the other hand, the estimate can be adjusted in the following sense. For every $1 \leq i \leq k$, let $0=\alpha_{i 0}, \alpha_{i 1}, \ldots, \alpha_{i h_{i}}$ be all the numbers of the form $\operatorname{ord}_{p_{i}}\left(d_{j}\right)$ in an increasing order (we delete the repeating occurences of a number, and add a number $0=\alpha_{i 0}$ if it does not occur). Then, repeating the arguments from the solution one can obtain that

$$
s \geq 1+\sum_{i=1}^{k} \sum_{j=1}^{h_{i}}\left(p^{\alpha_{j}-\alpha_{j-1}}-1\right) .
$$

Note that $p^{\alpha}-1 \geq \alpha(p-1)$, and the equality is achieved only for $\alpha=1$. Hence, for reaching the minimal number of the progressions, one should have $\alpha_{i, j}=j$ for all $i, j$. In other words, for each $1 \leq j \leq \alpha_{i}$, there should be an index $t$ such that $\operatorname{ord}_{p_{i}}\left(d_{t}\right)=j$.

Solution 2. We start with introducing some notation. For positive integer $r$, we denote $[r]=\{1,2, \ldots, r\}$. Next, we say that a set of progressions $\mathcal{P}=\left\{P_{1}, \ldots, P_{s}\right\}$ cover $\mathbb{Z}$ if each integer belongs to some of them; we say that this covering is minimal if no proper subset of $\mathcal{P}$ covers $\mathbb{Z}$. Obviously, each covering contains a minimal subcovering.

Next, for a minimal covering $\left\{P_{1}, \ldots, P_{s}\right\}$ and for every $1 \leq i \leq s$, let $d_{i}$ be the step of progression $P_{i}$, and $h_{i}$ be some number which is contained in $P_{i}$ but in none of the other progressions. We assume that $n>1$, otherwise the problem is trivial. This implies $d_{i}>1$, otherwise the progression $P_{i}$ covers all the numbers, and $n=1$.

We will prove a more general statement, namely the following
Claim. Assume that the progressions $P_{1}, \ldots, P_{s}$ and number $n=p_{1}^{\alpha_{1}} \ldots p_{k}^{\alpha_{k}}>1$ are chosen as in the problem statement. Moreover, choose some nonempty set of indices $I=\left\{i_{1}, \ldots, i_{t}\right\} \subseteq[k]$ and some positive integer $\beta_{i} \leq \alpha_{i}$ for every $i \in I$. Consider the set of indices

$$
T=\left\{j: 1 \leq j \leq s, \text { and } p_{i}^{\alpha_{i}-\beta_{i}+1} \mid d_{j} \text { for some } i \in I\right\} .
$$

Then

$$
\begin{equation*}
|T| \geq 1+\sum_{i \in I} \beta_{i}\left(p_{i}-1\right) \tag{2}
\end{equation*}
$$

Observe that the Claim for $I=[k]$ and $\beta_{i}=\alpha_{i}$ implies the problem statement, since the left-hand side in (2) is not greater than $s$. Hence, it suffices to prove the Claim.

1. First, we prove the Claim assuming that all $d_{j}$ 's are prime numbers. If for some $1 \leq i \leq k$ we have at least $p_{i}$ progressions with the step $p_{i}$, then they do not intersect and hence cover all the integers; it means that there are no other progressions, and $n=p_{i}$; the Claim is trivial in this case.

Now assume that for every $1 \leq i \leq k$, there are not more than $p_{i}-1$ progressions with step $p_{i}$; each such progression covers the numbers with a fixed residue modulo $p_{i}$, therefore there exists a residue $q_{i} \bmod p_{i}$ which is not touched by these progressions. By the Chinese Remainder Theorem, there exists a number $q$ such that $q \equiv q_{i}\left(\bmod p_{i}\right)$ for all $1 \leq i \leq k$; this number cannot be covered by any progression with step $p_{i}$, hence it is not covered at all. A contradiction.
2. Now, we assume that the general Claim is not valid, and hence we consider a counterexample $\left\{P_{1}, \ldots, P_{s}\right\}$ for the Claim; we can choose it to be minimal in the following sense:

- the number $n$ is minimal possible among all the counterexamples;
- the sum $\sum_{i} d_{i}$ is minimal possible among all the counterexamples having the chosen value of $n$.

As was mentioned above, not all numbers $d_{i}$ are primes; hence we can assume that $d_{1}$ is composite, say $p_{1} \mid d_{1}$ and $d_{1}^{\prime}=\frac{d_{1}}{p_{1}}>1$. Consider a progression $P_{1}^{\prime}$ having the step $d_{1}^{\prime}$, and containing $P_{1}$. We will focus on two coverings constructed as follows.
(i) Surely, the progressions $P_{1}^{\prime}, P_{2}, \ldots, P_{s}$ cover $\mathbb{Z}$, though this covering in not necessarily minimal. So, choose some minimal subcovering $\mathcal{P}^{\prime}$ in it; surely $P_{1}^{\prime} \in \mathcal{P}^{\prime}$ since $h_{1}$ is not covered by $P_{2}, \ldots, P_{s}$, so we may assume that $\mathcal{P}^{\prime}=\left\{P_{1}^{\prime}, P_{2}, \ldots, P_{s^{\prime}}\right\}$ for some $s^{\prime} \leq s$. Furthermore, the period of the covering $\mathcal{P}^{\prime}$ can appear to be less than $n$; so we denote this period by

$$
n^{\prime}=p_{1}^{\alpha_{1}-\sigma_{1}} \ldots p_{k}^{\alpha_{k}-\sigma_{k}}=\text { l.c.m. }\left(d_{1}^{\prime}, d_{2}, \ldots, d_{s^{\prime}}\right)
$$

Observe that for each $P_{j} \notin \mathcal{P}^{\prime}$, we have $h_{j} \in P_{1}^{\prime}$, otherwise $h_{j}$ would not be covered by $\mathcal{P}$.
(ii) On the other hand, each nonempty set of the form $R_{i}=P_{i} \cap P_{1}^{\prime}(1 \leq i \leq s)$ is also a progression with a step $r_{i}=$ l.c.m. $\left(d_{i}, d_{1}^{\prime}\right)$, and such sets cover $P_{1}^{\prime}$. Scaling these progressions with the ratio $1 / d_{1}^{\prime}$, we obtain the progressions $Q_{i}$ with steps $q_{i}=r_{i} / d_{1}^{\prime}$ which cover $\mathbb{Z}$. Now we choose a minimal subcovering $\mathcal{Q}$ of this covering; again we should have $Q_{1} \in \mathcal{Q}$ by the reasons of $h_{1}$. Now, denote the period of $\mathcal{Q}$ by

$$
n^{\prime \prime}=\text { l.c.m. }\left\{q_{i}: Q_{i} \in \mathcal{Q}\right\}=\frac{\text { l.c.m. }\left\{r_{i}: Q_{i} \in \mathcal{Q}\right\}}{d_{1}^{\prime}}=\frac{p_{1}^{\gamma_{1}} \ldots p_{k}^{\gamma_{k}}}{d_{1}^{\prime}} .
$$

Note that if $h_{j} \in P_{1}^{\prime}$, then the image of $h_{j}$ under the scaling can be covered by $Q_{j}$ only; so, in this case we have $Q_{j} \in \mathcal{Q}$.

Our aim is to find the desired number of progressions in coverings $\mathcal{P}$ and $\mathcal{Q}$. First, we have $n \geq n^{\prime}$, and the sum of the steps in $\mathcal{P}^{\prime}$ is less than that in $\mathcal{P}$; hence the Claim is valid for $\mathcal{P}^{\prime}$. We apply it to the set of indices $I^{\prime}=\left\{i \in I: \beta_{i}>\sigma_{i}\right\}$ and the exponents $\beta_{i}^{\prime}=\beta_{i}-\sigma_{i}$; hence the set under consideration is

$$
T^{\prime}=\left\{j: 1 \leq j \leq s^{\prime}, \text { and } p_{i}^{\left(\alpha_{i}-\sigma_{i}\right)-\beta_{i}^{\prime}+1}=p_{i}^{\alpha_{i}-\beta_{i}+1} \mid d_{j} \text { for some } i \in I^{\prime}\right\} \subseteq T \cap\left[s^{\prime}\right],
$$

and we obtain that

$$
\left|T \cap\left[s^{\prime}\right]\right| \geq\left|T^{\prime}\right| \geq 1+\sum_{i \in I^{\prime}}\left(\beta_{i}-\sigma_{i}\right)\left(p_{i}-1\right)=1+\sum_{i \in I}\left(\beta_{i}-\sigma_{i}\right)_{+}\left(p_{i}-1\right),
$$

where $(x)_{+}=\max \{x, 0\}$; the latter equality holds as for $i \notin I^{\prime}$ we have $\beta_{i} \leq \sigma_{i}$.
Observe that $x=(x-y)_{+}+\min \{x, y\}$ for all $x, y$. So, if we find at least

$$
G=\sum_{i \in I} \min \left\{\beta_{i}, \sigma_{i}\right\}\left(p_{i}-1\right)
$$

indices in $T \cap\left\{s^{\prime}+1, \ldots, s\right\}$, then we would have

$$
|T|=\left|T \cap\left[s^{\prime}\right]\right|+\left|T \cap\left\{s^{\prime}+1, \ldots, s\right\}\right| \geq 1+\sum_{i \in I}\left(\left(\beta_{i}-\sigma_{i}\right)_{+}+\min \left\{\beta_{i}, \sigma_{i}\right\}\right)\left(p_{i}-1\right)=1+\sum_{i \in I} \beta_{i}\left(p_{i}-1\right)
$$

thus leading to a contradiction with the choice of $\mathcal{P}$. We will find those indices among the indices of progressions in $\mathcal{Q}$.
3. Now denote $I^{\prime \prime}=\left\{i \in I: \sigma_{i}>0\right\}$ and consider some $i \in I^{\prime \prime}$; then $p_{i}^{\alpha_{i}} \nmid n^{\prime}$. On the other hand, there exists an index $j(i)$ such that $p_{i}^{\alpha_{i}} \mid d_{j(i)}$; this means that $d_{j(i)} \nmid n^{\prime}$ and hence $P_{j(i)}$ cannot appear in $\mathcal{P}^{\prime}$, so $j(i)>s^{\prime}$. Moreover, we have observed before that in this case $h_{j(i)} \in P_{1}^{\prime}$, hence $Q_{j(i)} \in \mathcal{Q}$. This means that $q_{j(i)} \mid n^{\prime \prime}$, therefore $\gamma_{i}=\alpha_{i}$ for each $i \in I^{\prime \prime}$ (recall here that $q_{i}=r_{i} / d_{1}^{\prime}$ and hence $\left.d_{j(i)}\left|r_{j(i)}\right| d_{1}^{\prime} n^{\prime \prime}\right)$.

Let $d_{1}^{\prime}=p_{1}^{\tau_{1}} \ldots p_{k}^{\tau_{k}}$. Then $n^{\prime \prime}=p_{1}^{\gamma_{1}-\tau_{1}} \ldots p_{k}^{\gamma_{i}-\tau_{i}}$. Now, if $i \in I^{\prime \prime}$, then for every $\beta$ the condition $p_{i}^{\left(\gamma_{i}-\tau_{i}\right)-\beta+1} \mid q_{j}$ is equivalent to $p_{i}^{\alpha_{i}-\beta+1} \mid r_{j}$.

Note that $n^{\prime \prime} \leq n / d_{1}^{\prime}<n$, hence we can apply the Claim to the covering $\mathcal{Q}$. We perform this with the set of indices $I^{\prime \prime}$ and the exponents $\beta_{i}^{\prime \prime}=\min \left\{\beta_{i}, \sigma_{i}\right\}>0$. So, the set under consideration is

$$
\begin{aligned}
T^{\prime \prime} & =\left\{j: Q_{j} \in \mathcal{Q}, \text { and } p_{i}^{\left(\gamma_{i}-\tau_{i}\right)-\min \left\{\beta_{i}, \sigma_{i}\right\}+1} \mid q_{j} \text { for some } i \in I^{\prime \prime}\right\} \\
& =\left\{j: Q_{j} \in \mathcal{Q}, \text { and } p_{i}^{\alpha_{i}-\min \left\{\beta_{i}, \sigma_{i}\right\}+1} \mid r_{j} \text { for some } i \in I^{\prime \prime}\right\},
\end{aligned}
$$

and we obtain $\left|T^{\prime \prime}\right| \geq 1+G$. Finally, we claim that $T^{\prime \prime} \subseteq T \cap\left(\{1\} \cup\left\{s^{\prime}+1, \ldots, s\right\}\right)$; then we will obtain $\left|T \cap\left\{s^{\prime}+1, \ldots, s\right\}\right| \geq G$, which is exactly what we need.

To prove this, consider any $j \in T^{\prime \prime}$. Observe first that $\alpha_{i}-\min \left\{\beta_{i}, \sigma_{i}\right\}+1>\alpha_{i}-\sigma_{i} \geq \tau_{i}$, hence from $p_{i}^{\alpha_{i}-\min \left\{\beta_{i}, \sigma_{i}\right\}+1} \mid r_{j}=$ l.c.m. $\left(d_{1}^{\prime}, d_{j}\right)$ we have $p_{i}^{\alpha_{i}-\min \left\{\beta_{i}, \sigma_{i}\right\}+1} \mid d_{j}$, which means that $j \in T$. Next, the exponent of $p_{i}$ in $d_{j}$ is greater than that in $n^{\prime}$, which means that $P_{j} \notin \mathcal{P}^{\prime}$. This may appear only if $j=1$ or $j>s^{\prime}$, as desired. This completes the proof.

Comment 2. A grid analogue of the Claim is also valid. It reads as following.
Claim. Assume that the grid $N$ is covered by subgrids $L_{1}, L_{2}, \ldots, L_{s}$ so that
(ii') each subgrid contains a point which is not covered by other subgrids;
(iii) for each coordinate axis, there exists a subgrid $L_{i}$ orthogonal to this axis.

Choose some set of indices $I=\left\{i_{1}, \ldots, i_{t}\right\} \subset[k]$, and consider the set of indices

$$
T=\left\{j: 1 \leq j \leq s, \text { and } L_{j} \text { is orthogonal to the } i \text { th axis for some } i \in I\right\} .
$$

Then

$$
|T| \geq 1+\sum_{i \in I}\left(n_{i}-1\right) .
$$

This Claim may be proved almost in the same way as in Solution 1.

## Geometry

G1. Let $A B C$ be an acute triangle with $D, E, F$ the feet of the altitudes lying on $B C, C A, A B$ respectively. One of the intersection points of the line $E F$ and the circumcircle is $P$. The lines $B P$ and $D F$ meet at point $Q$. Prove that $A P=A Q$.
(United Kingdom)
Solution 1. The line $E F$ intersects the circumcircle at two points. Depending on the choice of $P$, there are two different cases to consider.

Case 1: The point $P$ lies on the ray $E F$ (see Fig. 1).
Let $\angle C A B=\alpha, \angle A B C=\beta$ and $\angle B C A=\gamma$. The quadrilaterals $B C E F$ and $C A F D$ are cyclic due to the right angles at $D, E$ and $F$. So,

$$
\begin{aligned}
& \angle B D F=180^{\circ}-\angle F D C=\angle C A F=\alpha, \\
& \angle A F E=180^{\circ}-\angle E F B=\angle B C E=\gamma, \\
& \angle D F B=180^{\circ}-\angle A F D=\angle D C A=\gamma .
\end{aligned}
$$

Since $P$ lies on the arc $A B$ of the circumcircle, $\angle P B A<\angle B C A=\gamma$. Hence, we have

$$
\angle P B D+\angle B D F=\angle P B A+\angle A B D+\angle B D F<\gamma+\beta+\alpha=180^{\circ},
$$

and the point $Q$ must lie on the extensions of $B P$ and $D F$ beyond the points $P$ and $F$, respectively.

From the cyclic quadrilateral $A P B C$ we get

$$
\angle Q P A=180^{\circ}-\angle A P B=\angle B C A=\gamma=\angle D F B=\angle Q F A .
$$

Hence, the quadrilateral $A Q P F$ is cyclic. Then $\angle A Q P=180^{\circ}-\angle P F A=\angle A F E=\gamma$.
We obtained that $\angle A Q P=\angle Q P A=\gamma$, so the triangle $A Q P$ is isosceles, $A P=A Q$.


Fig. 1


Fig. 2

Case 2: The point $P$ lies on the ray $F E$ (see Fig. 2). In this case the point $Q$ lies inside the segment $F D$.

Similarly to the first case, we have

$$
\angle Q P A=\angle B C A=\gamma=\angle D F B=180^{\circ}-\angle A F Q
$$

Hence, the quadrilateral $A F Q P$ is cyclic.
Then $\angle A Q P=\angle A F P=\angle A F E=\gamma=\angle Q P A$. The triangle $A Q P$ is isosceles again, $\angle A Q P=\angle Q P A$ and thus $A P=A Q$.
Comment. Using signed angles, the two possible configurations can be handled simultaneously, without investigating the possible locations of $P$ and $Q$.

Solution 2. For arbitrary points $X, Y$ on the circumcircle, denote by $\widehat{X Y}$ the central angle of the arc $X Y$.

Let $P$ and $P^{\prime}$ be the two points where the line $E F$ meets the circumcircle; let $P$ lie on the arc $A B$ and let $P^{\prime}$ lie on the $\operatorname{arc} C A$. Let $B P$ and $B P^{\prime}$ meet the line $D F$ and $Q$ and $Q^{\prime}$, respectively (see Fig. 3). We will prove that $A P=A P^{\prime}=A Q=A Q^{\prime}$.


Fig. 3
Like in the first solution, we have $\angle A F E=\angle B F P=\angle D F B=\angle B C A=\gamma$ from the cyclic quadrilaterals $B C E F$ and $C A F D$.

By $\overparen{P B}+\overparen{P^{\prime} A}=2 \angle A F P^{\prime}=2 \gamma=2 \angle B C A=\overparen{A P}+\overparen{P B}$, we have

$$
\begin{equation*}
\widehat{A P}=\widetilde{P^{\prime} A}, \quad \angle P B A=\angle A B P^{\prime} \quad \text { and } \quad A P=A P^{\prime} \tag{1}
\end{equation*}
$$

Due to $\overparen{A P}=\overparen{P^{\prime} A}$, the lines $B P$ and $B Q^{\prime}$ are symmetrical about line $A B$.
Similarly, by $\angle B F P=\angle Q^{\prime} F B$, the lines $F P$ and $F Q^{\prime}$ are symmetrical about $A B$. It follows that also the points $P$ and $P^{\prime}$ are symmetrical to $Q^{\prime}$ and $Q$, respectively. Therefore,

$$
\begin{equation*}
A P=A Q^{\prime} \quad \text { and } \quad A P^{\prime}=A Q \tag{2}
\end{equation*}
$$

The relations (1) and (2) together prove $A P=A P^{\prime}=A Q=A Q^{\prime}$.

G2. Point $P$ lies inside triangle $A B C$. Lines $A P, B P, C P$ meet the circumcircle of $A B C$ again at points $K, L, M$, respectively. The tangent to the circumcircle at $C$ meets line $A B$ at $S$. Prove that $S C=S P$ if and only if $M K=M L$.

Solution 1. We assume that $C A>C B$, so point $S$ lies on the ray $A B$.
From the similar triangles $\triangle P K M \sim \triangle P C A$ and $\triangle P L M \sim \triangle P C B$ we get $\frac{P M}{K M}=\frac{P A}{C A}$ and $\frac{L M}{P M}=\frac{C B}{P B}$. Multiplying these two equalities, we get

$$
\frac{L M}{K M}=\frac{C B}{C A} \cdot \frac{P A}{P B}
$$

Hence, the relation $M K=M L$ is equivalent to $\frac{C B}{C A}=\frac{P B}{P A}$.
Denote by $E$ the foot of the bisector of angle $B$ in triangle $A B C$. Recall that the locus of points $X$ for which $\frac{X A}{X B}=\frac{C A}{C B}$ is the Apollonius circle $\Omega$ with the center $Q$ on the line $A B$, and this circle passes through $C$ and $E$. Hence, we have $M K=M L$ if and only if $P$ lies on $\Omega$, that is $Q P=Q C$.


Fig. 1

Now we prove that $S=Q$, thus establishing the problem statement. We have $\angle C E S=$ $\angle C A E+\angle A C E=\angle B C S+\angle E C B=\angle E C S$, so $S C=S E$. Hence, the point $S$ lies on $A B$ as well as on the perpendicular bisector of $C E$ and therefore coincides with $Q$.

Solution 2. As in the previous solution, we assume that $S$ lies on the ray $A B$.

1. Let $P$ be an arbitrary point inside both the circumcircle $\omega$ of the triangle $A B C$ and the angle $A S C$, the points $K, L, M$ defined as in the problem. We claim that $S P=S C$ implies $M K=M L$.

Let $E$ and $F$ be the points of intersection of the line $S P$ with $\omega$, point $E$ lying on the segment $S P$ (see Fig. 2).


Fig. 2

We have $S P^{2}=S C^{2}=S A \cdot S B$, so $\frac{S P}{S B}=\frac{S A}{S P}$, and hence $\triangle P S A \sim \triangle B S P$. Then $\angle B P S=\angle S A P$. Since $2 \angle B P S=\overparen{B E}+\overparen{L F}$ and $2 \angle S A P=\overparen{B E}+\overparen{E K}$ we have

$$
\begin{equation*}
\overparen{L F}=\overparen{E K} \tag{1}
\end{equation*}
$$

On the other hand, from $\angle S P C=\angle S C P$ we have $\overparen{E C}+\overparen{M F}=\widehat{E C}+\overparen{E M}$, or

$$
\begin{equation*}
\overparen{M F}=\overparen{E M} \tag{2}
\end{equation*}
$$

From (1) and (2) we get $\widehat{M F L}=\widehat{M F}+\overparen{F L}=\widehat{M E}+\widehat{E K}=\widehat{M E K}$ and hence $M K=M L$. The claim is proved.
2. We are left to prove the converse. So, assume that $M K=M L$, and introduce the points $E$ and $F$ as above. We have $S C^{2}=S E \cdot S F$; hence, there exists a point $P^{\prime}$ lying on the segment $E F$ such that $S P^{\prime}=S C$ (see Fig. 3).


Fig. 3

Assume that $P \neq P^{\prime}$. Let the lines $A P^{\prime}, B P^{\prime}, C P^{\prime}$ meet $\omega$ again at points $K^{\prime}, L^{\prime}, M^{\prime}$ respectively. Now, if $P^{\prime}$ lies on the segment $P F$ then by the first part of the solution we have $\widehat{M^{\prime} F L^{\prime}}=\widehat{M^{\prime} E K^{\prime}}$. On the other hand, we have $\widehat{M F L}>\widehat{M^{\prime} F L^{\prime}}=\widehat{M^{\prime} E K^{\prime}}>\widehat{M E K}$, therefore $\widehat{M F L}>\widehat{M E K}$ which contradicts $M K=M L$.

Similarly, if point $P^{\prime}$ lies on the segment $E P$ then we get $\widehat{M F L}<\widehat{M E K}$ which is impossible. Therefore, the points $P$ and $P^{\prime}$ coincide and hence $S P=S P^{\prime}=S C$.

Solution 3. We present a different proof of the converse direction, that is, $M K=M L \Rightarrow$ $S P=S C$. As in the previous solutions we assume that $C A>C B$, and the line $S P$ meets $\omega$ at $E$ and $F$.

From $M L=M K$ we get $\widehat{M E K}=\widehat{M F L}$. Now we claim that $\widehat{M E}=\widehat{M F}$ and $\widehat{E K}=\widehat{F L}$.
To the contrary, suppose first that $\widehat{M E}>\widehat{M F}$; then $\overparen{E K}=\widehat{M E K}-\overparen{M E}<\widehat{M F L}-\overparen{M F}=$ $\overparen{F L}$. Now, the inequality $\overparen{M E}>\overparen{M F}$ implies $2 \angle S C M=\overparen{E C}+\overparen{M E}>\overparen{E C}+\overparen{M F}=2 \angle S P C$ and hence $S P>S C$. On the other hand, the inequality $\overparen{E K}<\overparen{F L}$ implies $2 \angle S P K=$ $\overparen{E K}+\overparen{A F}<\overparen{F L}+\overparen{A F}=2 \angle A B L$, hence

$$
\angle S P A=180^{\circ}-\angle S P K>180^{\circ}-\angle A B L=\angle S B P
$$



Fig. 4
Consider the point $A^{\prime}$ on the ray $S A$ for which $\angle S P A^{\prime}=\angle S B P$; in our case, this point lies on the segment $S A$ (see Fig. 4). Then $\triangle S B P \sim \triangle S P A^{\prime}$ and $S P^{2}=S B \cdot S A^{\prime}<S B \cdot S A=S C^{2}$. Therefore, $S P<S C$ which contradicts $S P>S C$.

Similarly, one can prove that the inequality $\widehat{M E}<\overparen{M F}$ is also impossible. So, we get $\overparen{M E}=\overparen{M F}$ and therefore $2 \angle S C M=\widehat{E C}+\overparen{M E}=\overparen{E C}+\overparen{M F}=2 \angle S P C$, which implies $S C=S P$.

G3. Let $A_{1} A_{2} \ldots A_{n}$ be a convex polygon. Point $P$ inside this polygon is chosen so that its projections $P_{1}, \ldots, P_{n}$ onto lines $A_{1} A_{2}, \ldots, A_{n} A_{1}$ respectively lie on the sides of the polygon. Prove that for arbitrary points $X_{1}, \ldots, X_{n}$ on sides $A_{1} A_{2}, \ldots, A_{n} A_{1}$ respectively,

$$
\max \left\{\frac{X_{1} X_{2}}{P_{1} P_{2}}, \ldots, \frac{X_{n} X_{1}}{P_{n} P_{1}}\right\} \geq 1
$$

(Armenia)

Solution 1. Denote $P_{n+1}=P_{1}, X_{n+1}=X_{1}, A_{n+1}=A_{1}$.
Lemma. Let point $Q$ lies inside $A_{1} A_{2} \ldots A_{n}$. Then it is contained in at least one of the circumcircles of triangles $X_{1} A_{2} X_{2}, \ldots, X_{n} A_{1} X_{1}$.
Proof. If $Q$ lies in one of the triangles $X_{1} A_{2} X_{2}, \ldots, X_{n} A_{1} X_{1}$, the claim is obvious. Otherwise $Q$ lies inside the polygon $X_{1} X_{2} \ldots X_{n}$ (see Fig. 1). Then we have

$$
\begin{aligned}
& \left(\angle X_{1} A_{2} X_{2}+\angle X_{1} Q X_{2}\right)+\cdots+\left(\angle X_{n} A_{1} X_{1}+\angle X_{n} Q X_{1}\right) \\
& \quad=\left(\angle X_{1} A_{1} X_{2}+\cdots+\angle X_{n} A_{1} X_{1}\right)+\left(\angle X_{1} Q X_{2}+\cdots+\angle X_{n} Q X_{1}\right)=(n-2) \pi+2 \pi=n \pi
\end{aligned}
$$

hence there exists an index $i$ such that $\angle X_{i} A_{i+1} X_{i+1}+\angle X_{i} Q X_{i+1} \geq \frac{\pi n}{n}=\pi$. Since the quadrilateral $Q X_{i} A_{i+1} X_{i+1}$ is convex, this means exactly that $Q$ is contained the circumcircle of $\triangle X_{i} A_{i+1} X_{i+1}$, as desired.

Now we turn to the solution. Applying lemma, we get that $P$ lies inside the circumcircle of triangle $X_{i} A_{i+1} X_{i+1}$ for some $i$. Consider the circumcircles $\omega$ and $\Omega$ of triangles $P_{i} A_{i+1} P_{i+1}$ and $X_{i} A_{i+1} X_{i+1}$ respectively (see Fig. 2); let $r$ and $R$ be their radii. Then we get $2 r=A_{i+1} P \leq 2 R$ (since $P$ lies inside $\Omega$ ), hence

$$
P_{i} P_{i+1}=2 r \sin \angle P_{i} A_{i+1} P_{i+1} \leq 2 R \sin \angle X_{i} A_{i+1} X_{i+1}=X_{i} X_{i+1},
$$

QED.


Fig. 1


Fig. 2

Solution 2. As in Solution 1, we assume that all indices of points are considered modulo $n$.
We will prove a bit stronger inequality, namely

$$
\max \left\{\frac{X_{1} X_{2}}{P_{1} P_{2}} \cos \alpha_{1}, \ldots, \frac{X_{n} X_{1}}{P_{n} P_{1}} \cos \alpha_{n}\right\} \geq 1
$$

where $\alpha_{i}(1 \leq i \leq n)$ is the angle between lines $X_{i} X_{i+1}$ and $P_{i} P_{i+1}$. We denote $\beta_{i}=\angle A_{i} P_{i} P_{i-1}$ and $\gamma_{i}=\angle A_{i+1} P_{i} P_{i+1}$ for all $1 \leq i \leq n$.

Suppose that for some $1 \leq i \leq n$, point $X_{i}$ lies on the segment $A_{i} P_{i}$, while point $X_{i+1}$ lies on the segment $P_{i+1} A_{i+2}$. Then the projection of the segment $X_{i} X_{i+1}$ onto the line $P_{i} P_{i+1}$ contains segment $P_{i} P_{i+1}$, since $\gamma_{i}$ and $\beta_{i+1}$ are acute angles (see Fig. 3). Therefore, $X_{i} X_{i+1} \cos \alpha_{i} \geq$ $P_{i} P_{i+1}$, and in this case the statement is proved.

So, the only case left is when point $X_{i}$ lies on segment $P_{i} A_{i+1}$ for all $1 \leq i \leq n$ (the case when each $X_{i}$ lies on segment $A_{i} P_{i}$ is completely analogous).

Now, assume to the contrary that the inequality

$$
\begin{equation*}
X_{i} X_{i+1} \cos \alpha_{i}<P_{i} P_{i+1} \tag{1}
\end{equation*}
$$

holds for every $1 \leq i \leq n$. Let $Y_{i}$ and $Y_{i+1}^{\prime}$ be the projections of $X_{i}$ and $X_{i+1}$ onto $P_{i} P_{i+1}$. Then inequality (1) means exactly that $Y_{i} Y_{i+1}^{\prime}<P_{i} P_{i+1}$, or $P_{i} Y_{i}>P_{i+1} Y_{i+1}^{\prime}$ (again since $\gamma_{i}$ and $\beta_{i+1}$ are acute; see Fig. 4). Hence, we have

$$
X_{i} P_{i} \cos \gamma_{i}>X_{i+1} P_{i+1} \cos \beta_{i+1}, \quad 1 \leq i \leq n
$$

Multiplying these inequalities, we get

$$
\begin{equation*}
\cos \gamma_{1} \cos \gamma_{2} \cdots \cos \gamma_{n}>\cos \beta_{1} \cos \beta_{2} \cdots \cos \beta_{n} \tag{2}
\end{equation*}
$$

On the other hand, the sines theorem applied to triangle $P P_{i} P_{i+1}$ provides

$$
\frac{P P_{i}}{P P_{i+1}}=\frac{\sin \left(\frac{\pi}{2}-\beta_{i+1}\right)}{\sin \left(\frac{\pi}{2}-\gamma_{i}\right)}=\frac{\cos \beta_{i+1}}{\cos \gamma_{i}} .
$$

Multiplying these equalities we get

$$
1=\frac{\cos \beta_{2}}{\cos \gamma_{1}} \cdot \frac{\cos \beta_{3}}{\cos \gamma_{2}} \cdots \frac{\cos \beta_{1}}{\cos \gamma_{n}}
$$

which contradicts (2).


Fig. 3
Fig. 4

G4. Let $I$ be the incenter of a triangle $A B C$ and $\Gamma$ be its circumcircle. Let the line $A I$ intersect $\Gamma$ at a point $D \neq A$. Let $F$ and $E$ be points on side $B C$ and arc $B D C$ respectively such that $\angle B A F=\angle C A E<\frac{1}{2} \angle B A C$. Finally, let $G$ be the midpoint of the segment IF. Prove that the lines $D G$ and $E I$ intersect on $\Gamma$.
(Hong Kong)
Solution 1. Let $X$ be the second point of intersection of line $E I$ with $\Gamma$, and $L$ be the foot of the bisector of angle $B A C$. Let $G^{\prime}$ and $T$ be the points of intersection of segment $D X$ with lines $I F$ and $A F$, respectively. We are to prove that $G=G^{\prime}$, or $I G^{\prime}=G^{\prime} F$. By the Menelaus theorem applied to triangle $A I F$ and line $D X$, it means that we need the relation

$$
1=\frac{G^{\prime} F}{I G^{\prime}}=\frac{T F}{A T} \cdot \frac{A D}{I D}, \quad \text { or } \quad \frac{T F}{A T}=\frac{I D}{A D} .
$$

Let the line $A F$ intersect $\Gamma$ at point $K \neq A$ (see Fig. 1); since $\angle B A K=\angle C A E$ we have $\widehat{B K}=\overparen{C E}$, hence $K E \| B C$. Notice that $\angle I A T=\angle D A K=\angle E A D=\angle E X D=\angle I X T$, so the points $I, A, X, T$ are concyclic. Hence we have $\angle I T A=\angle I X A=\angle E X A=\angle E K A$, so $I T\|K E\| B C$. Therefore we obtain $\frac{T F}{A T}=\frac{I L}{A I}$.

Since $C I$ is the bisector of $\angle A C L$, we get $\frac{I L}{A I}=\frac{C L}{A C}$. Furthermore, $\angle D C L=\angle D C B=$ $\angle D A B=\angle C A D=\frac{1}{2} \angle B A C$, hence the triangles $D C L$ and $D A C$ are similar; therefore we get $\frac{C L}{A C}=\frac{D C}{A D}$. Finally, it is known that the midpoint $D$ of $\operatorname{arc} B C$ is equidistant from points $I$, $B, C$, hence $\frac{D C}{A D}=\frac{I D}{A D}$.

Summarizing all these equalities, we get

$$
\frac{T F}{A T}=\frac{I L}{A I}=\frac{C L}{A C}=\frac{D C}{A D}=\frac{I D}{A D}
$$

as desired.


Fig. 1


Fig. 2

Comment. The equality $\frac{A I}{I L}=\frac{A D}{D I}$ is known and can be obtained in many different ways. For instance, one can consider the inversion with center $D$ and radius $D C=D I$. This inversion takes $\widehat{B A C}$ to the segment $B C$, so point $A$ goes to $L$. Hence $\frac{I L}{D I}=\frac{A I}{A D}$, which is the desired equality.

Solution 2. As in the previous solution, we introduce the points $X, T$ and $K$ and note that it suffice to prove the equality

$$
\frac{T F}{A T}=\frac{D I}{A D} \quad \Longleftrightarrow \quad \frac{T F+A T}{A T}=\frac{D I+A D}{A D} \quad \Longleftrightarrow \quad \frac{A T}{A D}=\frac{A F}{D I+A D}
$$

Since $\angle F A D=\angle E A I$ and $\angle T D A=\angle X D A=\angle X E A=\angle I E A$, we get that the triangles $A T D$ and $A I E$ are similar, therefore $\frac{A T}{A D}=\frac{A I}{A E}$.

Next, we also use the relation $D B=D C=D I$. Let $J$ be the point on the extension of segment $A D$ over point $D$ such that $D J=D I=D C$ (see Fig. 2). Then $\angle D J C=$ $\angle J C D=\frac{1}{2}(\pi-\angle J D C)=\frac{1}{2} \angle A D C=\frac{1}{2} \angle A B C=\angle A B I$. Moreover, $\angle B A I=\angle J A C$, hence triangles $A B I$ and $A J C$ are similar, so $\frac{A B}{A J}=\frac{A I}{A C}$, or $A B \cdot A C=A J \cdot A I=(D I+A D) \cdot A I$.

On the other hand, we get $\angle A B F=\angle A B C=\angle A E C$ and $\angle B A F=\angle C A E$, so triangles $A B F$ and $A E C$ are also similar, which implies $\frac{A F}{A C}=\frac{A B}{A E}$, or $A B \cdot A C=A F \cdot A E$.

Summarizing we get

$$
(D I+A D) \cdot A I=A B \cdot A C=A F \cdot A E \quad \Rightarrow \quad \frac{A I}{A E}=\frac{A F}{A D+D I} \quad \Rightarrow \quad \frac{A T}{A D}=\frac{A F}{A D+D I}
$$

as desired.
Comment. In fact, point $J$ is an excenter of triangle $A B C$.

G5. Let $A B C D E$ be a convex pentagon such that $B C \| A E, A B=B C+A E$, and $\angle A B C=$ $\angle C D E$. Let $M$ be the midpoint of $C E$, and let $O$ be the circumcenter of triangle $B C D$. Given that $\angle D M O=90^{\circ}$, prove that $2 \angle B D A=\angle C D E$.
(Ukraine)
Solution 1. Choose point $T$ on ray $A E$ such that $A T=A B$; then from $A E \| B C$ we have $\angle C B T=\angle A T B=\angle A B T$, so $B T$ is the bisector of $\angle A B C$. On the other hand, we have $E T=A T-A E=A B-A E=B C$, hence quadrilateral $B C T E$ is a parallelogram, and the midpoint $M$ of its diagonal $C E$ is also the midpoint of the other diagonal $B T$.

Next, let point $K$ be symmetrical to $D$ with respect to $M$. Then $O M$ is the perpendicular bisector of segment $D K$, and hence $O D=O K$, which means that point $K$ lies on the circumcircle of triangle $B C D$. Hence we have $\angle B D C=\angle B K C$. On the other hand, the angles $B K C$ and $T D E$ are symmetrical with respect to $M$, so $\angle T D E=\angle B K C=\angle B D C$.

Therefore, $\angle B D T=\angle B D E+\angle E D T=\angle B D E+\angle B D C=\angle C D E=\angle A B C=180^{\circ}-$ $\angle B A T$. This means that the points $A, B, D, T$ are concyclic, and hence $\angle A D B=\angle A T B=$ $\frac{1}{2} \angle A B C=\frac{1}{2} \angle C D E$, as desired.


Solution 2. Let $\angle C B D=\alpha, \angle B D C=\beta, \angle A D E=\gamma$, and $\angle A B C=\angle C D E=2 \varphi$. Then we have $\angle A D B=2 \varphi-\beta-\gamma, \angle B C D=180^{\circ}-\alpha-\beta, \angle A E D=360^{\circ}-\angle B C D-\angle C D E=$ $180^{\circ}-2 \varphi+\alpha+\beta$, and finally $\angle D A E=180^{\circ}-\angle A D E-\angle A E D=2 \varphi-\alpha-\beta-\gamma$.


Let $N$ be the midpoint of $C D$; then $\angle D N O=90^{\circ}=\angle D M O$, hence points $M, N$ lie on the circle with diameter $O D$. Now, if points $O$ and $M$ lie on the same side of $C D$, we have $\angle D M N=\angle D O N=\frac{1}{2} \angle D O C=\alpha$; in the other case, we have $\angle D M N=180^{\circ}-\angle D O N=\alpha ;$
so, in both cases $\angle D M N=\alpha$ (see Figures). Next, since $M N$ is a midline in triangle $C D E$, we have $\angle M D E=\angle D M N=\alpha$ and $\angle N D M=2 \varphi-\alpha$.

Now we apply the sine rule to the triangles $A B D, A D E$ (twice), $B C D$ and $M N D$ obtaining

$$
\begin{gathered}
\frac{A B}{A D}=\frac{\sin (2 \varphi-\beta-\gamma)}{\sin (2 \varphi-\alpha)}, \quad \frac{A E}{A D}=\frac{\sin \gamma}{\sin (2 \varphi-\alpha-\beta)}, \quad \frac{D E}{A D}=\frac{\sin (2 \varphi-\alpha-\beta-\gamma)}{\sin (2 \varphi-\alpha-\beta)} \\
\frac{B C}{C D}=\frac{\sin \beta}{\sin \alpha}, \quad \frac{C D}{D E}=\frac{C D / 2}{D E / 2}=\frac{N D}{N M}=\frac{\sin \alpha}{\sin (2 \varphi-\alpha)}
\end{gathered}
$$

which implies

$$
\frac{B C}{A D}=\frac{B C}{C D} \cdot \frac{C D}{D E} \cdot \frac{D E}{A D}=\frac{\sin \beta \cdot \sin (2 \varphi-\alpha-\beta-\gamma)}{\sin (2 \varphi-\alpha) \cdot \sin (2 \varphi-\alpha-\beta)}
$$

Hence, the condition $A B=A E+B C$, or equivalently $\frac{A B}{A D}=\frac{A E+B C}{A D}$, after multiplying by the common denominator rewrites as

$$
\begin{gathered}
\quad \sin (2 \varphi-\alpha-\beta) \cdot \sin (2 \varphi-\beta-\gamma)=\sin \gamma \cdot \sin (2 \varphi-\alpha)+\sin \beta \cdot \sin (2 \varphi-\alpha-\beta-\gamma) \\
\Longleftrightarrow \cos (\gamma-\alpha)-\cos (4 \varphi-2 \beta-\alpha-\gamma)=\cos (2 \varphi-\alpha-2 \beta-\gamma)-\cos (2 \varphi+\gamma-\alpha) \\
\Longleftrightarrow \cos (\gamma-\alpha)+\cos (2 \varphi+\gamma-\alpha)=\cos (2 \varphi-\alpha-2 \beta-\gamma)+\cos (4 \varphi-2 \beta-\alpha-\gamma) \\
\Longleftrightarrow \cos \varphi \cdot \cos (\varphi+\gamma-\alpha)=\cos \varphi \cdot \cos (3 \varphi-2 \beta-\alpha-\gamma) \\
\Longleftrightarrow \cos \varphi \cdot(\cos (\varphi+\gamma-\alpha)-\cos (3 \varphi-2 \beta-\alpha-\gamma))=0 \\
\Longleftrightarrow \cos \varphi \cdot \sin (2 \varphi-\beta-\alpha) \cdot \sin (\varphi-\beta-\gamma)=0 .
\end{gathered}
$$

Since $2 \varphi-\beta-\alpha=180^{\circ}-\angle A E D<180^{\circ}$ and $\varphi=\frac{1}{2} \angle A B C<90^{\circ}$, it follows that $\varphi=\beta+\gamma$, hence $\angle B D A=2 \varphi-\beta-\gamma=\varphi=\frac{1}{2} \angle C D E$, as desired.

G6. The vertices $X, Y, Z$ of an equilateral triangle $X Y Z$ lie respectively on the sides $B C$, $C A, A B$ of an acute-angled triangle $A B C$. Prove that the incenter of triangle $A B C$ lies inside triangle $X Y Z$.

G6 ${ }^{\prime}$. The vertices $X, Y, Z$ of an equilateral triangle $X Y Z$ lie respectively on the sides $B C, C A, A B$ of a triangle $A B C$. Prove that if the incenter of triangle $A B C$ lies outside triangle $X Y Z$, then one of the angles of triangle $A B C$ is greater than $120^{\circ}$.
(Bulgaria)
Solution 1 for G6. We will prove a stronger fact; namely, we will show that the incenter $I$ of triangle $A B C$ lies inside the incircle of triangle $X Y Z$ (and hence surely inside triangle $X Y Z$ itself). We denote by $d(U, V W)$ the distance between point $U$ and line $V W$.

Denote by $O$ the incenter of $\triangle X Y Z$ and by $r, r^{\prime}$ and $R^{\prime}$ the inradii of triangles $A B C, X Y Z$ and the circumradius of $X Y Z$, respectively. Then we have $R^{\prime}=2 r^{\prime}$, and the desired inequality is $O I \leq r^{\prime}$. We assume that $O \neq I$; otherwise the claim is trivial.

Let the incircle of $\triangle A B C$ touch its sides $B C, A C, A B$ at points $A_{1}, B_{1}, C_{1}$ respectively. The lines $I A_{1}, I B_{1}, I C_{1}$ cut the plane into 6 acute angles, each one containing one of the points $A_{1}, B_{1}, C_{1}$ on its border. We may assume that $O$ lies in an angle defined by lines $I A_{1}$, $I C_{1}$ and containing point $C_{1}$ (see Fig. 1). Let $A^{\prime}$ and $C^{\prime}$ be the projections of $O$ onto lines $I A_{1}$ and $I C_{1}$, respectively.

Since $O X=R^{\prime}$, we have $d(O, B C) \leq R^{\prime}$. Since $O A^{\prime} \| B C$, it follows that $d\left(A^{\prime}, B C\right)=$ $A^{\prime} I+r \leq R^{\prime}$, or $A^{\prime} I \leq R^{\prime}-r$. On the other hand, the incircle of $\triangle X Y Z$ lies inside $\triangle A B C$, hence $d(O, A B) \geq r^{\prime}$, and analogously we get $d(O, A B)=C^{\prime} C_{1}=r-I C^{\prime} \geq r^{\prime}$, or $I C^{\prime} \leq r-r^{\prime}$.


Fig. 1


Fig. 2

Finally, the quadrilateral $I A^{\prime} O C^{\prime}$ is circumscribed due to the right angles at $A^{\prime}$ and $C^{\prime}$ (see Fig. 2). On its circumcircle, we have $\widehat{A^{\prime} O C^{\prime}}=2 \angle A^{\prime} I C^{\prime}<180^{\circ}=\widehat{O C^{\prime} I}$, hence $180^{\circ} \geq$ $\widetilde{I C^{\prime}}>\widetilde{A^{\prime} O}$. This means that $I C^{\prime}>A^{\prime} O$. Finally, we have $O I \leq I A^{\prime}+A^{\prime} O<I A^{\prime}+I C^{\prime} \leq$ $\left(R^{\prime}-r\right)+\left(r-r^{\prime}\right)=R^{\prime}-r^{\prime}=r^{\prime}$, as desired.

Solution 2 for G6. Assume the contrary. Then the incenter $I$ should lie in one of triangles $A Y Z, B X Z, C X Y$ - assume that it lies in $\triangle A Y Z$. Let the incircle $\omega$ of $\triangle A B C$ touch sides $B C, A C$ at point $A_{1}, B_{1}$ respectively. Without loss of generality, assume that point $A_{1}$ lies on segment $C X$. In this case we will show that $\angle C>90^{\circ}$ thus leading to a contradiction.

Note that $\omega$ intersects each of the segments $X Y$ and $Y Z$ at two points; let $U, U^{\prime}$ and $V$, $V^{\prime}$ be the points of intersection of $\omega$ with $X Y$ and $Y Z$, respectively $\left(U Y>U^{\prime} Y, V Y>V^{\prime} Y\right.$; see Figs. 3 and 4). Note that $60^{\circ}=\angle X Y Z=\frac{1}{2}\left(\overparen{U V}-\overparen{U^{\prime} V^{\prime}}\right) \leq \frac{1}{2} \overparen{U V}$, hence $\overparen{U V} \geq 120^{\circ}$.

On the other hand, since $I$ lies in $\triangle A Y Z$, we get $\widehat{V U V^{\prime}}<180^{\circ}$, hence $\widehat{U A_{1} U^{\prime}} \leq \widehat{U A_{1} V^{\prime}}<$ $180^{\circ}-\overparen{U V} \leq 60^{\circ}$.

Now, two cases are possible due to the order of points $Y, B_{1}$ on segment $A C$.


Fig. 3


Fig. 4

Case 1. Let point $Y$ lie on the segment $A B_{1}$ (see Fig. 3). Then we have $\angle Y X C=$ $\frac{1}{2}\left(\widehat{A_{1} U^{\prime}}-\widehat{A_{1} U}\right) \leq \frac{1}{2} \widehat{U A_{1} U^{\prime}}<30^{\circ}$; analogously, we get $\angle X Y C \leq \frac{1}{2} \widehat{U A_{1} U^{\prime}}<30^{\circ}$. Therefore, $\angle Y C X=180^{\circ}-\angle Y X C-\angle X Y C>120^{\circ}$, as desired.

Case 2. Now let point $Y$ lie on the segment $C B_{1}$ (see Fig. 4). Analogously, we obtain $\angle Y X C<30^{\circ}$. Next, $\angle I Y X>\angle Z Y X=60^{\circ}$, but $\angle I Y X<\angle I Y B_{1}$, since $Y B_{1}$ is a tangent and $Y X$ is a secant line to circle $\omega$ from point $Y$. Hence, we get $120^{\circ}<\angle I Y B_{1}+\angle I Y X=$ $\angle B_{1} Y X=\angle Y X C+\angle Y C X<30^{\circ}+\angle Y C X$, hence $\angle Y C X>120^{\circ}-30^{\circ}=90^{\circ}$, as desired.

Comment. In the same way, one can prove a more general
Claim. Let the vertices $X, Y, Z$ of a triangle $X Y Z$ lie respectively on the sides $B C, C A, A B$ of a triangle $A B C$. Suppose that the incenter of triangle $A B C$ lies outside triangle $X Y Z$, and $\alpha$ is the least angle of $\triangle X Y Z$. Then one of the angles of triangle $A B C$ is greater than $3 \alpha-90^{\circ}$.

Solution for G6'. Assume the contrary. As in Solution 2, we assume that the incenter $I$ of $\triangle A B C$ lies in $\triangle A Y Z$, and the tangency point $A_{1}$ of $\omega$ and $B C$ lies on segment $C X$. Surely, $\angle Y Z A \leq 180^{\circ}-\angle Y Z X=120^{\circ}$, hence points $I$ and $Y$ lie on one side of the perpendicular bisector to $X Y$; therefore $I X>I Y$. Moreover, $\omega$ intersects segment $X Y$ at two points, and therefore the projection $M$ of $I$ onto $X Y$ lies on the segment $X Y$. In this case, we will prove that $\angle C>120^{\circ}$.

Let $Y K, Y L$ be two tangents from point $Y$ to $\omega$ (points $K$ and $A_{1}$ lie on one side of $X Y$; if $Y$ lies on $\omega$, we say $K=L=Y$ ); one of the points $K$ and $L$ is in fact a tangency point $B_{1}$ of $\omega$ and $A C$. From symmetry, we have $\angle Y I K=\angle Y I L$. On the other hand, since $I X>I Y$, we get $X M<X Y$ which implies $\angle A_{1} X Y<\angle K Y X$.

Next, we have $\angle M I Y=90^{\circ}-\angle I Y X<90^{\circ}-\angle Z Y X=30^{\circ}$. Since $I A_{1} \perp A_{1} X, I M \perp X Y$, $I K \perp Y K$ we get $\angle M I A_{1}=\angle A_{1} X Y<\angle K Y X=\angle M I K$. Finally, we get

$$
\begin{aligned}
\angle A_{1} I K<\angle A_{1} I L=( & \left.\angle A_{1} I M+\angle M I K\right)+(\angle K I Y+\angle Y I L) \\
& <2 \angle M I K+2 \angle K I Y=2 \angle M I Y<60^{\circ} .
\end{aligned}
$$

Hence, $\angle A_{1} I B_{1}<60^{\circ}$, and therefore $\angle A C B=180^{\circ}-\angle A_{1} I B_{1}>120^{\circ}$, as desired.


Fig. 5


Fig. 6

Comment 1. The estimate claimed in $\mathrm{G}^{\prime}$ is sharp. Actually, if $\angle B A C>120^{\circ}$, one can consider an equilateral triangle $X Y Z$ with $Z=A, Y \in A C, X \in B C$ (such triangle exists since $\angle A C B<60^{\circ}$ ). It intersects with the angle bisector of $\angle B A C$ only at point $A$, hence it does not contain $I$.

Comment 2. As in the previous solution, there is a generalization for an arbitrary triangle $X Y Z$, but here we need some additional condition. The statement reads as follows.
Claim. Let the vertices $X, Y, Z$ of a triangle $X Y Z$ lie respectively on the sides $B C, C A, A B$ of a triangle $A B C$. Suppose that the incenter of triangle $A B C$ lies outside triangle $X Y Z, \alpha$ is the least angle of $\triangle X Y Z$, and all sides of triangle $X Y Z$ are greater than $2 r \cot \alpha$, where $r$ is the inradius of $\triangle A B C$. Then one of the angles of triangle $A B C$ is greater than $2 \alpha$.

The additional condition is needed to verify that $X M>Y M$ since it cannot be shown in the original way. Actually, we have $\angle M Y I>\alpha, I M<r$, hence $Y M<r \cot \alpha$. Now, if we have $X Y=X M+Y M>2 r \cot \alpha$, then surely $X M>Y M$.

On the other hand, this additional condition follows easily from the conditions of the original problem. Actually, if $I \in \triangle A Y Z$, then the diameter of $\omega$ parallel to $Y Z$ is contained in $\triangle A Y Z$ and is thus shorter than $Y Z$. Hence $Y Z>2 r>2 r \cot 60^{\circ}$.

G7. Three circular arcs $\gamma_{1}, \gamma_{2}$, and $\gamma_{3}$ connect the points $A$ and $C$. These arcs lie in the same half-plane defined by line $A C$ in such a way that arc $\gamma_{2}$ lies between the arcs $\gamma_{1}$ and $\gamma_{3}$. Point $B$ lies on the segment $A C$. Let $h_{1}, h_{2}$, and $h_{3}$ be three rays starting at $B$, lying in the same half-plane, $h_{2}$ being between $h_{1}$ and $h_{3}$. For $i, j=1,2,3$, denote by $V_{i j}$ the point of intersection of $h_{i}$ and $\gamma_{j}$ (see the Figure below).

Denote by $\widehat{V_{i j} V_{k j}} \sqrt{V_{k \ell} V_{i \ell}}$ the curved quadrilateral, whose sides are the segments $V_{i j} V_{i \ell}, V_{k j} V_{k \ell}$ and $\operatorname{arcs} V_{i j} V_{k j}$ and $V_{i \ell} V_{k \ell}$. We say that this quadrilateral is circumscribed if there exists a circle touching these two segments and two arcs.

Prove that if the curved quadrilaterals $\sqrt{V_{11} V_{21}} \sqrt{V_{22} V_{12}}, \sqrt{12 V_{22}} \sqrt{23} V_{13}, \sqrt{21 V_{31}} \sqrt{V_{32} V_{22}}$ are circumscribed, then the curved quadrilateral $\widehat{V_{22} V_{32}} \sqrt[V_{33} V_{23}]{ }$ is circumscribed, too.


Fig. 1

Solution. Denote by $O_{i}$ and $R_{i}$ the center and the radius of $\gamma_{i}$, respectively. Denote also by $H$ the half-plane defined by $A C$ which contains the whole configuration. For every point $P$ in the half-plane $H$, denote by $d(P)$ the distance between $P$ and line $A C$. Furthermore, for any $r>0$, denote by $\Omega(P, r)$ the circle with center $P$ and radius $r$.
Lemma 1. For every $1 \leq i<j \leq 3$, consider those circles $\Omega(P, r)$ in the half-plane $H$ which are tangent to $h_{i}$ and $h_{j}$.
(a) The locus of the centers of these circles is the angle bisector $\beta_{i j}$ between $h_{i}$ and $h_{j}$.
(b) There is a constant $u_{i j}$ such that $r=u_{i j} \cdot d(P)$ for all such circles.

Proof. Part (a) is obvious. To prove part (b), notice that the circles which are tangent to $h_{i}$ and $h_{j}$ are homothetic with the common homothety center $B$ (see Fig. 2). Then part (b) also becomes trivial.

Lemma 2. For every $1 \leq i<j \leq 3$, consider those circles $\Omega(P, r)$ in the half-plane $H$ which are externally tangent to $\gamma_{i}$ and internally tangent to $\gamma_{j}$.
(a) The locus of the centers of these circles is an ellipse arc $\varepsilon_{i j}$ with end-points $A$ and $C$.
(b) There is a constant $v_{i j}$ such that $r=v_{i j} \cdot d(P)$ for all such circles.

Proof. (a) Notice that the circle $\Omega(P, r)$ is externally tangent to $\gamma_{i}$ and internally tangent to $\gamma_{j}$ if and only if $O_{i} P=R_{i}+r$ and $O_{j}=R_{j}-r$. Therefore, for each such circle we have

$$
O_{i} P+O_{j} P=O_{i} A+O_{j} A=O_{i} C+O_{j} C=R_{i}+R_{j}
$$

Such points lie on an ellipse with foci $O_{i}$ and $O_{j}$; the diameter of this ellipse is $R_{i}+R_{j}$, and it passes through the points $A$ and $C$. Let $\varepsilon_{i j}$ be that arc $A C$ of the ellipse which runs inside the half plane $H$ (see Fig. 3.)

This ellipse arc lies between the arcs $\gamma_{i}$ and $\gamma_{j}$. Therefore, if some point $P$ lies on $\varepsilon_{i j}$, then $O_{i} P>R_{i}$ and $O_{j} P<R_{j}$. Now, we choose $r=O_{i} P-R_{i}=R_{j}-O_{j} P>0$; then the

circle $\Omega(P, r)$ touches $\gamma_{i}$ externally and touches $\gamma_{j}$ internally, so $P$ belongs to the locus under investigation.
(b) Let $\vec{\rho}=\overrightarrow{A P}, \vec{\rho}_{i}=\overrightarrow{A O_{i}}$, and $\vec{\rho}_{j}=\overrightarrow{A O_{j}}$; let $d_{i j}=O_{i} O_{j}$, and let $\vec{v}$ be a unit vector orthogonal to $A C$ and directed toward $H$. Then we have $\left|\vec{\rho}_{i}\right|=R_{i},\left|\vec{\rho}_{j}\right|=R_{j},\left|\overrightarrow{O_{i} P}\right|=$ $\left|\vec{\rho}-\vec{\rho}_{i}\right|=R_{i}+r,\left|\overrightarrow{O_{j} P}\right|=\left|\vec{\rho}-\vec{\rho}_{j}\right|=R_{j}-r$, hence

$$
\begin{gathered}
\left(\vec{\rho}-\vec{\rho}_{i}\right)^{2}-\left(\vec{\rho}-\vec{\rho}_{j}\right)^{2}=\left(R_{i}+r\right)^{2}-\left(R_{j}-r\right)^{2}, \\
\left(\vec{\rho}_{i}^{2}-\vec{\rho}_{j}^{2}\right)+2 \vec{\rho} \cdot\left(\vec{\rho}_{j}-\vec{\rho}_{i}\right)=\left(R_{i}^{2}-R_{j}^{2}\right)+2 r\left(R_{i}+R_{j}\right), \\
d_{i j} \cdot d(P)=d_{i j} \vec{v} \cdot \vec{\rho}=\left(\vec{\rho}_{j}-\vec{\rho}_{i}\right) \cdot \vec{\rho}=r\left(R_{i}+R_{j}\right) .
\end{gathered}
$$

Therefore,

$$
r=\frac{d_{i j}}{R_{i}+R_{j}} \cdot d(P)
$$

and the value $v_{i j}=\frac{d_{i j}}{R_{i}+R_{j}}$ does not depend on $P$.
Lemma 3. The curved quadrilateral $\mathcal{Q}_{i j}=\sqrt{i, j V_{i+1, j}} V_{i+1, j+1} V_{i, j+1}$ is circumscribed if and only if $u_{i, i+1}=v_{j, j+1}$.
Proof. First suppose that the curved quadrilateral $\mathcal{Q}_{i j}$ is circumscribed and $\Omega(P, r)$ is its inscribed circle. By Lemma 1 and Lemma 2 we have $r=u_{i, i+1} \cdot d(P)$ and $r=v_{j, j+1} \cdot d(P)$ as well. Hence, $u_{i, i+1}=v_{j, j+1}$.

To prove the opposite direction, suppose $u_{i, i+1}=v_{j, j+1}$. Let $P$ be the intersection of the angle bisector $\beta_{i, i+1}$ and the ellipse arc $\varepsilon_{j, j+1}$. Choose $r=u_{i, i+1} \cdot d(P)=v_{j, j+1} \cdot d(P)$. Then the circle $\Omega(P, r)$ is tangent to the half lines $h_{i}$ and $h_{i+1}$ by Lemma 1 , and it is tangent to the $\operatorname{arcs} \gamma_{j}$ and $\gamma_{j+1}$ by Lemma 2. Hence, the curved quadrilateral $\mathcal{Q}_{i j}$ is circumscribed.

By Lemma 3, the statement of the problem can be reformulated to an obvious fact: If the equalities $u_{12}=v_{12}, u_{12}=v_{23}$, and $u_{23}=v_{12}$ hold, then $u_{23}=v_{23}$ holds as well.

Comment 1. Lemma 2(b) (together with the easy Lemma 1(b)) is the key tool in this solution. If one finds this fact, then the solution can be finished in many ways. That is, one can find a circle touching three of $h_{2}, h_{3}, \gamma_{2}$, and $\gamma_{3}$, and then prove that it is tangent to the fourth one in either synthetic or analytical way. Both approaches can be successful.

Here we present some discussion about this key Lemma.

1. In the solution above we chose an analytic proof for Lemma 2(b) because we expect that most students will use coordinates or vectors to examine the locus of the centers, and these approaches are less case-sensitive.

Here we outline a synthetic proof. We consider only the case when $P$ does not lie in the line $O_{i} O_{j}$. The other case can be obtained as a limit case, or computed in a direct way.

Let $S$ be the internal homothety center between the circles of $\gamma_{i}$ and $\gamma_{j}$, lying on $O_{i} O_{j}$; this point does not depend on $P$. Let $U$ and $V$ be the points of tangency of circle $\sigma=\Omega(P, r)$ with $\gamma_{i}$ and $\gamma_{j}$, respectively (then $r=P U=P V$ ); in other words, points $U$ and $V$ are the intersection points of rays $O_{i} P, O_{j} P$ with arcs $\gamma_{i}, \gamma_{j}$ respectively (see Fig. 4).

Due to the theorem on three homothety centers (or just to the Menelaus theorem applied to triangle $O_{i} O_{j} P$ ), the points $U, V$ and $S$ are collinear. Let $T$ be the intersection point of line $A C$ and the common tangent to $\sigma$ and $\gamma_{i}$ at $U$; then $T$ is the radical center of $\sigma, \gamma_{i}$ and $\gamma_{j}$, hence $T V$ is the common tangent to $\sigma$ and $\gamma_{j}$.

Let $Q$ be the projection of $P$ onto the line $A C$. By the right angles, the points $U, V$ and $Q$ lie on the circle with diameter $P T$. From this fact and the equality $P U=P V$ we get $\angle U Q P=\angle U V P=$ $\angle V U P=\angle S U O_{i}$. Since $O_{i} S \| P Q$, we have $\angle S O_{i} U=\angle Q P U$. Hence, the triangles $S O_{i} U$ and $U P Q$ are similar and thus $\frac{r}{d(P)}=\frac{P U}{P Q}=\frac{O_{i} S}{O_{i} U}=\frac{O_{i} S}{R_{i}}$; the last expression is constant since $S$ is a constant point.


Fig. 4


Fig. 5
2. Using some known facts about conics, the same statement can be proved in a very short way. Denote by $\ell$ the directrix of ellipse of $\varepsilon_{i j}$ related to the focus $O_{j}$; since $\varepsilon_{i j}$ is symmetrical about $O_{i} O_{j}$, we have $\ell \| A C$. Recall that for each point $P \in \varepsilon_{i j}$, we have $P O_{j}=\epsilon \cdot d_{\ell}(P)$, where $d_{\ell}(P)$ is the distance from $P$ to $\ell$, and $\epsilon$ is the eccentricity of $\varepsilon_{i j}$ (see Fig. 5).

Now we have

$$
r=R_{j}-\left(R_{j}-r\right)=A O_{j}-P O_{j}=\epsilon\left(d_{\ell}(A)-d_{\ell}(P)\right)=\epsilon(d(P)-d(A))=\epsilon \cdot d(P)
$$

and $\epsilon$ does not depend on $P$.

Comment 2. One can find a spatial interpretations of the problem and the solution.
For every point $(x, y)$ and radius $r>0$, represent the circle $\Omega((x, y), r)$ by the point $(x, y, r)$ in space. This point is the apex of the cone with base circle $\Omega((x, y), r)$ and height $r$. According to Lemma 1 , the circles which are tangent to $h_{i}$ and $h_{j}$ correspond to the points of a half line $\beta_{i j}^{\prime}$, starting at $B$.

Now we translate Lemma 2. Take some $1 \leq i<j \leq 3$, and consider those circles which are internally tangent to $\gamma_{j}$. It is easy to see that the locus of the points which represent these circles is a subset of a cone, containing $\gamma_{j}$. Similarly, the circles which are externally tangent to $\gamma_{i}$ correspond to the points on the extension of another cone, which has its apex on the opposite side of the base plane $\Pi$. (See Fig. 6; for this illustration, the $z$-coordinates were multiplied by 2.)

The two cones are symmetric to each other (they have the same aperture, and their axes are parallel). As is well-known, it follows that the common points of the two cones are co-planar. So the intersection of the two cones is a a conic section - which is an ellipse, according to Lemma 2(a). The points which represent the circles touching $\gamma_{i}$ and $\gamma_{j}$ is an ellipse arc $\varepsilon_{i j}^{\prime}$ with end-points $A$ and $C$.


Fig. 6


Fig. 7

Thus, the curved quadrilateral $\mathcal{Q}_{i j}$ is circumscribed if and only if $\beta_{i, i+1}^{\prime}$ and $\varepsilon_{j, j+1}^{\prime}$ intersect, i.e. if they are coplanar. If three of the four curved quadrilaterals are circumscribed, it means that $\varepsilon_{12}^{\prime}, \varepsilon_{23}^{\prime}$, $\beta_{12}^{\prime}$ and $\beta_{23}^{\prime}$ lie in the same plane $\Sigma$, and the fourth intersection comes to existence, too (see Fig. 7).


A connection between mathematics and real life: the Palace of Creativity "Shabyt" ("Inspiration") in Astana

## Number Theory

N1. Find the least positive integer $n$ for which there exists a set $\left\{s_{1}, s_{2}, \ldots, s_{n}\right\}$ consisting of $n$ distinct positive integers such that

$$
\left(1-\frac{1}{s_{1}}\right)\left(1-\frac{1}{s_{2}}\right) \ldots\left(1-\frac{1}{s_{n}}\right)=\frac{51}{2010} .
$$

$\mathbf{N 1}^{\prime}$. Same as Problem N1, but the constant $\frac{51}{2010}$ is replaced by $\frac{42}{2010}$.
(Canada)
Answer for Problem N1. $n=39$.
Solution for Problem N1. Suppose that for some $n$ there exist the desired numbers; we may assume that $s_{1}<s_{2}<\cdots<s_{n}$. Surely $s_{1}>1$ since otherwise $1-\frac{1}{s_{1}}=0$. So we have $2 \leq s_{1} \leq s_{2}-1 \leq \cdots \leq s_{n}-(n-1)$, hence $s_{i} \geq i+1$ for each $i=1, \ldots, n$. Therefore

$$
\begin{aligned}
\frac{51}{2010} & =\left(1-\frac{1}{s_{1}}\right)\left(1-\frac{1}{s_{2}}\right) \ldots\left(1-\frac{1}{s_{n}}\right) \\
& \geq\left(1-\frac{1}{2}\right)\left(1-\frac{1}{3}\right) \ldots\left(1-\frac{1}{n+1}\right)=\frac{1}{2} \cdot \frac{2}{3} \cdots \frac{n}{n+1}=\frac{1}{n+1}
\end{aligned}
$$

which implies

$$
n+1 \geq \frac{2010}{51}=\frac{670}{17}>39
$$

so $n \geq 39$.
Now we are left to show that $n=39$ fits. Consider the set $\{2,3, \ldots, 33,35,36, \ldots, 40,67\}$ which contains exactly 39 numbers. We have

$$
\begin{equation*}
\frac{1}{2} \cdot \frac{2}{3} \cdots \frac{32}{33} \cdot \frac{34}{35} \cdots \frac{39}{40} \cdot \frac{66}{67}=\frac{1}{33} \cdot \frac{34}{40} \cdot \frac{66}{67}=\frac{17}{670}=\frac{51}{2010} \tag{1}
\end{equation*}
$$

hence for $n=39$ there exists a desired example.
Comment. One can show that the example (1) is unique.
Answer for Problem N1'. $n=48$.
Solution for Problem N1'. Suppose that for some $n$ there exist the desired numbers. In the same way we obtain that $s_{i} \geq i+1$. Moreover, since the denominator of the fraction $\frac{42}{2010}=\frac{7}{335}$ is divisible by 67 , some of $s_{i}$ 's should be divisible by 67 , so $s_{n} \geq s_{i} \geq 67$. This means that

$$
\frac{42}{2010} \geq \frac{1}{2} \cdot \frac{2}{3} \cdots \frac{n-1}{n} \cdot\left(1-\frac{1}{67}\right)=\frac{66}{67 n},
$$

which implies

$$
n \geq \frac{2010 \cdot 66}{42 \cdot 67}=\frac{330}{7}>47
$$

so $n \geq 48$.
Now we are left to show that $n=48$ fits. Consider the set $\{2,3, \ldots, 33,36,37, \ldots, 50,67\}$ which contains exactly 48 numbers. We have

$$
\frac{1}{2} \cdot \frac{2}{3} \cdots \frac{32}{33} \cdot \frac{35}{36} \cdots \frac{49}{50} \cdot \frac{66}{67}=\frac{1}{33} \cdot \frac{35}{50} \cdot \frac{66}{67}=\frac{7}{335}=\frac{42}{2010}
$$

hence for $n=48$ there exists a desired example.
Comment 1. In this version of the problem, the estimate needs one more step, hence it is a bit harder. On the other hand, the example in this version is not unique. Another example is

$$
\frac{1}{2} \cdot \frac{2}{3} \cdots \frac{46}{47} \cdot \frac{66}{67} \cdot \frac{329}{330}=\frac{1}{67} \cdot \frac{66}{330} \cdot \frac{329}{47}=\frac{7}{67 \cdot 5}=\frac{42}{2010} .
$$

Comment 2. N1' was the Proposer's formulation of the problem. We propose N1 according to the number of current IMO.

N2. Find all pairs $(m, n)$ of nonnegative integers for which

$$
\begin{equation*}
m^{2}+2 \cdot 3^{n}=m\left(2^{n+1}-1\right) \tag{1}
\end{equation*}
$$

(Australia)
Answer. $(6,3),(9,3),(9,5),(54,5)$.
Solution. For fixed values of $n$, the equation (1) is a simple quadratic equation in $m$. For $n \leq 5$ the solutions are listed in the following table.

| case | equation | discriminant | integer roots |
| :--- | :--- | :--- | :--- |
| $n=0$ | $m^{2}-m+2=0$ | -7 | none |
| $n=1$ | $m^{2}-3 m+6=0$ | -15 | none |
| $n=2$ | $m^{2}-7 m+18=0$ | -23 | none |
| $n=3$ | $m^{2}-15 m+54=0$ | 9 | $m=6$ and $m=9$ |
| $n=4$ | $m^{2}-31 m+162=0$ | 313 | none |
| $n=5$ | $m^{2}-63 m+486=0$ | $2025=45^{2}$ | $m=9$ and $m=54$ |

We prove that there is no solution for $n \geq 6$.
Suppose that ( $m, n$ ) satisfies (1) and $n \geq 6$. Since $m \mid 2 \cdot 3^{n}=m\left(2^{n+1}-1\right)-m^{2}$, we have $m=3^{p}$ with some $0 \leq p \leq n$ or $m=2 \cdot 3^{q}$ with some $0 \leq q \leq n$.

In the first case, let $q=n-p$; then

$$
2^{n+1}-1=m+\frac{2 \cdot 3^{n}}{m}=3^{p}+2 \cdot 3^{q}
$$

In the second case let $p=n-q$. Then

$$
2^{n+1}-1=m+\frac{2 \cdot 3^{n}}{m}=2 \cdot 3^{q}+3^{p}
$$

Hence, in both cases we need to find the nonnegative integer solutions of

$$
\begin{equation*}
3^{p}+2 \cdot 3^{q}=2^{n+1}-1, \quad p+q=n \tag{2}
\end{equation*}
$$

Next, we prove bounds for $p, q$. From (2) we get

$$
3^{p}<2^{n+1}=8^{\frac{n+1}{3}}<9^{\frac{n+1}{3}}=3^{\frac{2(n+1)}{3}}
$$

and

$$
2 \cdot 3^{q}<2^{n+1}=2 \cdot 8^{\frac{n}{3}}<2 \cdot 9^{\frac{n}{3}}=2 \cdot 3^{\frac{2 n}{3}}<2 \cdot 3^{\frac{2(n+1)}{3}}
$$

so $p, q<\frac{2(n+1)}{3}$. Combining these inequalities with $p+q=n$, we obtain

$$
\begin{equation*}
\frac{n-2}{3}<p, q<\frac{2(n+1)}{3} \tag{3}
\end{equation*}
$$

Now let $h=\min (p, q)$. By (3) we have $h>\frac{n-2}{3}$; in particular, we have $h>1$. On the left-hand side of (2), both terms are divisible by $3^{h}$, therefore $9\left|3^{h}\right| 2^{n+1}-1$. It is easy check that $\operatorname{ord}_{9}(2)=6$, so $9 \mid 2^{n+1}-1$ if and only if $6 \mid n+1$. Therefore, $n+1=6 r$ for some positive integer $r$, and we can write

$$
\begin{equation*}
2^{n+1}-1=4^{3 r}-1=\left(4^{2 r}+4^{r}+1\right)\left(2^{r}-1\right)\left(2^{r}+1\right) \tag{4}
\end{equation*}
$$

Notice that the factor $4^{2 r}+4^{r}+1=\left(4^{r}-1\right)^{2}+3 \cdot 4^{r}$ is divisible by 3 , but it is never divisible by 9 . The other two factors in (4), $2^{r}-1$ and $2^{r}+1$ are coprime: both are odd and their difference is 2 . Since the whole product is divisible by $3^{h}$, we have either $3^{h-1} \mid 2^{r}-1$ or $3^{h-1} \mid 2^{r}+1$. In any case, we have $3^{h-1} \leq 2^{r}+1$. Then

$$
\begin{gathered}
3^{h-1} \leq 2^{r}+1 \leq 3^{r}=3^{\frac{n+1}{6}} \\
\frac{n-2}{3}-1<h-1 \leq \frac{n+1}{6} \\
n<11
\end{gathered}
$$

But this is impossible since we assumed $n \geq 6$, and we proved $6 \mid n+1$.

N3. Find the smallest number $n$ such that there exist polynomials $f_{1}, f_{2}, \ldots, f_{n}$ with rational coefficients satisfying

$$
x^{2}+7=f_{1}(x)^{2}+f_{2}(x)^{2}+\cdots+f_{n}(x)^{2} .
$$

(Poland)
Answer. The smallest $n$ is 5 .
Solution 1. The equality $x^{2}+7=x^{2}+2^{2}+1^{2}+1^{2}+1^{2}$ shows that $n \leq 5$. It remains to show that $x^{2}+7$ is not a sum of four (or less) squares of polynomials with rational coefficients.

Suppose by way of contradiction that $x^{2}+7=f_{1}(x)^{2}+f_{2}(x)^{2}+f_{3}(x)^{2}+f_{4}(x)^{2}$, where the coefficients of polynomials $f_{1}, f_{2}, f_{3}$ and $f_{4}$ are rational (some of these polynomials may be zero).

Clearly, the degrees of $f_{1}, f_{2}, f_{3}$ and $f_{4}$ are at most 1 . Thus $f_{i}(x)=a_{i} x+b_{i}$ for $i=1,2,3,4$ and some rationals $a_{1}, b_{1}, a_{2}, b_{2}, a_{3}, b_{3}, a_{4}, b_{4}$. It follows that $x^{2}+7=\sum_{i=1}^{4}\left(a_{i} x+b_{i}\right)^{2}$ and hence

$$
\begin{equation*}
\sum_{i=1}^{4} a_{i}^{2}=1, \quad \sum_{i=1}^{4} a_{i} b_{i}=0, \quad \sum_{i=1}^{4} b_{i}^{2}=7 . \tag{1}
\end{equation*}
$$

Let $p_{i}=a_{i}+b_{i}$ and $q_{i}=a_{i}-b_{i}$ for $i=1,2,3,4$. Then

$$
\begin{aligned}
\sum_{i=1}^{4} p_{i}^{2} & =\sum_{i=1}^{4} a_{i}^{2}+2 \sum_{i=1}^{4} a_{i} b_{i}+\sum_{i=1}^{4} b_{i}^{2}=8, \\
\sum_{i=1}^{4} q_{i}^{2} & =\sum_{i=1}^{4} a_{i}^{2}-2 \sum_{i=1}^{4} a_{i} b_{i}+\sum_{i=1}^{4} b_{i}^{2}=8 \\
\text { and } \quad \sum_{i=1}^{4} p_{i} q_{i} & =\sum_{i=1}^{4} a_{i}^{2}-\sum_{i=1}^{4} b_{i}^{2}=-6,
\end{aligned}
$$

which means that there exist a solution in integers $x_{1}, y_{1}, x_{2}, y_{2}, x_{3}, y_{3}, x_{4}, y_{4}$ and $m>0$ of the system of equations
(i) $\sum_{i=1}^{4} x_{i}^{2}=8 m^{2}$,
(ii) $\sum_{i=1}^{4} y_{i}^{2}=8 m^{2}$,
(iii) $\sum_{i=1}^{4} x_{i} y_{i}=-6 m^{2}$.

We will show that such a solution does not exist.
Assume the contrary and consider a solution with minimal $m$. Note that if an integer $x$ is odd then $x^{2} \equiv 1(\bmod 8)$. Otherwise (i.e., if $x$ is even) we have $x^{2} \equiv 0(\bmod 8)$ or $x^{2} \equiv 4$ $(\bmod 8)$. Hence, by (i), we get that $x_{1}, x_{2}, x_{3}$ and $x_{4}$ are even. Similarly, by (ii), we get that $y_{1}, y_{2}, y_{3}$ and $y_{4}$ are even. Thus the LHS of (iii) is divisible by 4 and $m$ is also even. It follows that $\left(\frac{x_{1}}{2}, \frac{y_{1}}{2}, \frac{x_{2}}{2}, \frac{y_{2}}{2}, \frac{x_{3}}{2}, \frac{y_{3}}{2}, \frac{x_{4}}{2}, \frac{y_{4}}{2}, \frac{m}{2}\right)$ is a solution of the system of equations (i), (ii) and (iii), which contradicts the minimality of $m$.

Solution 2. We prove that $n \leq 4$ is impossible. Define the numbers $a_{i}, b_{i}$ for $i=1,2,3,4$ as in the previous solution.

By Euler's identity we have

$$
\begin{aligned}
\left(a_{1}^{2}+a_{2}^{2}+a_{3}^{2}+a_{4}^{2}\right)\left(b_{1}^{2}+b_{2}^{2}+b_{3}^{2}+b_{4}^{2}\right) & =\left(a_{1} b_{1}+a_{2} b_{2}+a_{3} b_{3}+a_{4} b_{4}\right)^{2}+\left(a_{1} b_{2}-a_{2} b_{1}+a_{3} b_{4}-a_{4} b_{3}\right)^{2} \\
& +\left(a_{1} b_{3}-a_{3} b_{1}+a_{4} b_{2}-a_{2} b_{4}\right)^{2}+\left(a_{1} b_{4}-a_{4} b_{1}+a_{2} b_{3}-a_{3} b_{2}\right)^{2} .
\end{aligned}
$$

So, using the relations (1) from the Solution 1 we get that

$$
\begin{equation*}
7=\left(\frac{m_{1}}{m}\right)^{2}+\left(\frac{m_{2}}{m}\right)^{2}+\left(\frac{m_{3}}{m}\right)^{2} \tag{2}
\end{equation*}
$$

where

$$
\begin{aligned}
& \frac{m_{1}}{m}=a_{1} b_{2}-a_{2} b_{1}+a_{3} b_{4}-a_{4} b_{3}, \\
& \frac{m_{2}}{m}=a_{1} b_{3}-a_{3} b_{1}+a_{4} b_{2}-a_{2} b_{4}, \\
& \frac{m_{3}}{m}=a_{1} b_{4}-a_{4} b_{1}+a_{2} b_{3}-a_{3} b_{2}
\end{aligned}
$$

and $m_{1}, m_{2}, m_{3} \in \mathbb{Z}, m \in \mathbb{N}$.
Let $m$ be a minimum positive integer number for which (2) holds. Then

$$
8 m^{2}=m_{1}^{2}+m_{2}^{2}+m_{3}^{2}+m^{2} .
$$

As in the previous solution, we get that $m_{1}, m_{2}, m_{3}, m$ are all even numbers. Then $\left(\frac{m_{1}}{2}, \frac{m_{2}}{2}, \frac{m_{3}}{2}, \frac{m}{2}\right)$ is also a solution of (2) which contradicts the minimality of $m$. So, we have $n \geq 5$. The example with $n=5$ is already shown in Solution 1 .

N4. Let $a, b$ be integers, and let $P(x)=a x^{3}+b x$. For any positive integer $n$ we say that the pair $(a, b)$ is $n$-good if $n \mid P(m)-P(k)$ implies $n \mid m-k$ for all integers $m, k$. We say that $(a, b)$ is very good if $(a, b)$ is $n$-good for infinitely many positive integers $n$.
(a) Find a pair $(a, b)$ which is 51 -good, but not very good.
(b) Show that all 2010-good pairs are very good.
(Turkey)
Solution. (a) We show that the pair $\left(1,-51^{2}\right)$ is good but not very good. Let $P(x)=x^{3}-51^{2} x$. Since $P(51)=P(0)$, the pair $\left(1,-51^{2}\right)$ is not $n$-good for any positive integer that does not divide 51 . Therefore, $\left(1,-51^{2}\right)$ is not very good.

On the other hand, if $P(m) \equiv P(k)(\bmod 51)$, then $m^{3} \equiv k^{3}(\bmod 51)$. By Fermat's theorem, from this we obtain

$$
m \equiv m^{3} \equiv k^{3} \equiv k \quad(\bmod 3) \quad \text { and } \quad m \equiv m^{33} \equiv k^{33} \equiv k \quad(\bmod 17)
$$

Hence we have $m \equiv k(\bmod 51)$. Therefore $\left(1,-51^{2}\right)$ is 51 -good.
(b) We will show that if a pair $(a, b)$ is 2010-good then $(a, b)$ is $67^{i}$-good for all positive integer $i$.
Claim 1. If $(a, b)$ is 2010 -good then $(a, b)$ is 67 -good.
Proof. Assume that $P(m)=P(k)(\bmod 67)$. Since 67 and 30 are coprime, there exist integers $m^{\prime}$ and $k^{\prime}$ such that $k^{\prime} \equiv k(\bmod 67), k^{\prime} \equiv 0(\bmod 30)$, and $m^{\prime} \equiv m(\bmod 67), m^{\prime} \equiv 0$ $(\bmod 30)$. Then we have $P\left(m^{\prime}\right) \equiv P(0) \equiv P\left(k^{\prime}\right)(\bmod 30)$ and $P\left(m^{\prime}\right) \equiv P(m) \equiv P(k) \equiv P\left(k^{\prime}\right)$ $(\bmod 67)$, hence $P\left(m^{\prime}\right) \equiv P\left(k^{\prime}\right)(\bmod 2010)$. This implies $m^{\prime} \equiv k^{\prime}(\bmod 2010)$ as $(a, b)$ is 2010-good. It follows that $m \equiv m^{\prime} \equiv k^{\prime} \equiv k(\bmod 67)$. Therefore, $(a, b)$ is 67 -good.
Claim 2. If $(a, b)$ is 67 -good then $67 \mid a$.
Proof. Suppose that $67 \nmid a$. Consider the sets $\left\{a t^{2}(\bmod 67): 0 \leq t \leq 33\right\}$ and $\left\{-3 a s^{2}-b\right.$ $\bmod 67: 0 \leq s \leq 33\}$. Since $a \not \equiv 0(\bmod 67)$, each of these sets has 34 elements. Hence they have at least one element in common. If $a t^{2} \equiv-3 a s^{2}-b(\bmod 67)$ then for $m=t \pm s, k=\mp 2 s$ we have

$$
\begin{aligned}
P(m)-P(k)=a\left(m^{3}-k^{3}\right)+b(m-k) & =(m-k)\left(a\left(m^{2}+m k+k^{2}\right)+b\right) \\
& =(t \pm 3 s)\left(a t^{2}+3 a s^{2}+b\right) \equiv 0 \quad(\bmod 67)
\end{aligned}
$$

Since $(a, b)$ is 67 -good, we must have $m \equiv k(\bmod 67)$ in both cases, that is, $t \equiv 3 s(\bmod 67)$ and $t \equiv-3 s(\bmod 67)$. This means $t \equiv s \equiv 0(\bmod 67)$ and $b \equiv-3 a s^{2}-a t^{2} \equiv 0(\bmod 67)$. But then $67 \mid P(7)-P(2)=67 \cdot 5 a+5 b$ and $67 \nmid 7-2$, contradicting that $(a, b)$ is 67 -good.
Claim 3. If $(a, b)$ is 2010-good then $(a, b)$ is $67^{i}$-good all $i \geq 1$.
Proof. By Claim 2, we have $67 \mid a$. If $67 \mid b$, then $P(x) \equiv P(0)(\bmod 67)$ for all $x$, contradicting that $(a, b)$ is 67 -good. Hence, $67 \nmid b$.

Suppose that $67^{i} \mid P(m)-P(k)=(m-k)\left(a\left(m^{2}+m k+k^{2}\right)+b\right)$. Since $67 \mid a$ and $67 \nmid b$, the second factor $a\left(m^{2}+m k+k^{2}\right)+b$ is coprime to 67 and hence $67^{i} \mid m-k$. Therefore, $(a, b)$ is $67^{i}$-good.
Comment 1. In the proof of Claim 2, the following reasoning can also be used. Since 3 is not a quadratic residue modulo 67 , either $a u^{2} \equiv-b(\bmod 67)$ or $3 a v^{2} \equiv-b(\bmod 67)$ has a solution. The settings $(m, k)=(u, 0)$ in the first case and $(m, k)=(v,-2 v)$ in the second case lead to $b \equiv 0$ $(\bmod 67)$.
Comment 2. The pair $(67,30)$ is $n$-good if and only if $n=d \cdot 67^{i}$, where $d \mid 30$ and $i \geq 0$. It shows that in part (b), one should deal with the large powers of 67 to reach the solution. The key property of number 67 is that it has the form $3 k+1$, so there exists a nontrivial cubic root of unity modulo 67 .

N5. Let $\mathbb{N}$ be the set of all positive integers. Find all functions $f: \mathbb{N} \rightarrow \mathbb{N}$ such that the number $(f(m)+n)(m+f(n))$ is a square for all $m, n \in \mathbb{N}$.
(U.S.A.)

Answer. All functions of the form $f(n)=n+c$, where $c \in \mathbb{N} \cup\{0\}$.
Solution. First, it is clear that all functions of the form $f(n)=n+c$ with a constant nonnegative integer $c$ satisfy the problem conditions since $(f(m)+n)(f(n)+m)=(n+m+c)^{2}$ is a square.

We are left to prove that there are no other functions. We start with the following Lemma. Suppose that $p \mid f(k)-f(\ell)$ for some prime $p$ and positive integers $k, \ell$. Then $p \mid k-\ell$. Proof. Suppose first that $p^{2} \mid f(k)-f(\ell)$, so $f(\ell)=f(k)+p^{2} a$ for some integer $a$. Take some positive integer $D>\max \{f(k), f(\ell)\}$ which is not divisible by $p$ and set $n=p D-f(k)$. Then the positive numbers $n+f(k)=p D$ and $n+f(\ell)=p D+(f(\ell)-f(k))=p(D+p a)$ are both divisible by $p$ but not by $p^{2}$. Now, applying the problem conditions, we get that both the numbers $(f(k)+n)(f(n)+k)$ and $(f(\ell)+n)(f(n)+\ell)$ are squares divisible by $p$ (and thus by $p^{2}$ ); this means that the multipliers $f(n)+k$ and $f(n)+\ell$ are also divisible by $p$, therefore $p \mid(f(n)+k)-(f(n)+\ell)=k-\ell$ as well.

On the other hand, if $f(k)-f(\ell)$ is divisible by $p$ but not by $p^{2}$, then choose the same number $D$ and set $n=p^{3} D-f(k)$. Then the positive numbers $f(k)+n=p^{3} D$ and $f(\ell)+n=$ $p^{3} D+(f(\ell)-f(k))$ are respectively divisible by $p^{3}$ (but not by $p^{4}$ ) and by $p$ (but not by $p^{2}$ ). Hence in analogous way we obtain that the numbers $f(n)+k$ and $f(n)+\ell$ are divisible by $p$, therefore $p \mid(f(n)+k)-(f(n)+\ell)=k-\ell$.

We turn to the problem. First, suppose that $f(k)=f(\ell)$ for some $k, \ell \in \mathbb{N}$. Then by Lemma we have that $k-\ell$ is divisible by every prime number, so $k-\ell=0$, or $k=\ell$. Therefore, the function $f$ is injective.

Next, consider the numbers $f(k)$ and $f(k+1)$. Since the number $(k+1)-k=1$ has no prime divisors, by Lemma the same holds for $f(k+1)-f(k)$; thus $|f(k+1)-f(k)|=1$.

Now, let $f(2)-f(1)=q,|q|=1$. Then we prove by induction that $f(n)=f(1)+q(n-1)$. The base for $n=1,2$ holds by the definition of $q$. For the step, if $n>1$ we have $f(n+1)=$ $f(n) \pm q=f(1)+q(n-1) \pm q$. Since $f(n) \neq f(n-2)=f(1)+q(n-2)$, we get $f(n)=f(1)+q n$, as desired.

Finally, we have $f(n)=f(1)+q(n-1)$. Then $q$ cannot be -1 since otherwise for $n \geq f(1)+1$ we have $f(n) \leq 0$ which is impossible. Hence $q=1$ and $f(n)=(f(1)-1)+n$ for each $n \in \mathbb{N}$, and $f(1)-1 \geq 0$, as desired.

N6. The rows and columns of a $2^{n} \times 2^{n}$ table are numbered from 0 to $2^{n}-1$. The cells of the table have been colored with the following property being satisfied: for each $0 \leq i, j \leq 2^{n}-1$, the $j$ th cell in the $i$ th row and the $(i+j)$ th cell in the $j$ th row have the same color. (The indices of the cells in a row are considered modulo $2^{n}$.)

Prove that the maximal possible number of colors is $2^{n}$.

Solution. Throughout the solution we denote the cells of the table by coordinate pairs; $(i, j)$ refers to the $j$ th cell in the $i$ th row.

Consider the directed graph, whose vertices are the cells of the board, and the edges are the arrows $(i, j) \rightarrow(j, i+j)$ for all $0 \leq i, j \leq 2^{n}-1$. From each vertex $(i, j)$, exactly one edge passes $\left(\right.$ to $\left(j, i+j \bmod 2^{n}\right)$ ); conversely, to each cell $(j, k)$ exactly one edge is directed (from the cell $\left.\left(k-j \bmod 2^{n}, j\right)\right)$. Hence, the graph splits into cycles.

Now, in any coloring considered, the vertices of each cycle should have the same color by the problem condition. On the other hand, if each cycle has its own color, the obtained coloring obviously satisfies the problem conditions. Thus, the maximal possible number of colors is the same as the number of cycles, and we have to prove that this number is $2^{n}$.

Next, consider any cycle $\left(i_{1}, j_{1}\right),\left(i_{2}, j_{2}\right), \ldots$; we will describe it in other terms. Define a sequence $\left(a_{0}, a_{1}, \ldots\right)$ by the relations $a_{0}=i_{1}, a_{1}=j_{1}, a_{n+1}=a_{n}+a_{n-1}$ for all $n \geq 1$ (we say that such a sequence is a Fibonacci-type sequence). Then an obvious induction shows that $i_{k} \equiv a_{k-1}\left(\bmod 2^{n}\right), j_{k} \equiv a_{k}\left(\bmod 2^{n}\right)$. Hence we need to investigate the behavior of Fibonacci-type sequences modulo $2^{n}$.

Denote by $F_{0}, F_{1}, \ldots$ the Fibonacci numbers defined by $F_{0}=0, F_{1}=1$, and $F_{n+2}=$ $F_{n+1}+F_{n}$ for $n \geq 0$. We also set $F_{-1}=1$ according to the recurrence relation.

For every positive integer $m$, denote by $\nu(m)$ the exponent of 2 in the prime factorization of $m$, i.e. for which $2^{\nu(m)} \mid m$ but $2^{\nu(m)+1} \backslash m$.
Lemma 1. For every Fibonacci-type sequence $a_{0}, a_{1}, a_{2}, \ldots$, and every $k \geq 0$, we have $a_{k}=$ $F_{k-1} a_{0}+F_{k} a_{1}$.
Proof. Apply induction on $k$. The base cases $k=0,1$ are trivial. For the step, from the induction hypothesis we get

$$
a_{k+1}=a_{k}+a_{k-1}=\left(F_{k-1} a_{0}+F_{k} a_{1}\right)+\left(F_{k-2} a_{0}+F_{k-1} a_{1}\right)=F_{k} a_{0}+F_{k+1} a_{1} .
$$

Lemma 2. For every $m \geq 3$,
(a) we have $\nu\left(F_{3 \cdot 2^{m-2}}\right)=m$;
(b) $d=3 \cdot 2^{m-2}$ is the least positive index for which $2^{m} \mid F_{d}$;
(c) $F_{3 \cdot 2^{m-2}+1} \equiv 1+2^{m-1}\left(\bmod 2^{m}\right)$.

Proof. Apply induction on $m$. In the base case $m=3$ we have $\nu\left(F_{3 \cdot 2^{m-2}}\right)=F_{6}=8$, so $\nu\left(F_{3 \cdot 2^{m-2}}\right)=\nu(8)=3$, the preceding Fibonacci-numbers are not divisible by 8, and indeed $F_{3 \cdot 2^{m-2}+1}=F_{7}=13 \equiv 1+4(\bmod 8)$.

Now suppose that $m>3$ and let $k=3 \cdot 2^{m-3}$. By applying Lemma 1 to the Fibonacci-type sequence $F_{k}, F_{k+1}, \ldots$ we get

$$
\begin{gathered}
F_{2 k}=F_{k-1} F_{k}+F_{k} F_{k+1}=\left(F_{k+1}-F_{k}\right) F_{k}+F_{k+1} F_{k}=2 F_{k+1} F_{k}-F_{k}^{2}, \\
F_{2 k+1}=F_{k} \cdot F_{k}+F_{k+1} \cdot F_{k+1}=F_{k}^{2}+F_{k+1}^{2} .
\end{gathered}
$$

By the induction hypothesis, $\nu\left(F_{k}\right)=m-1$, and $F_{k+1}$ is odd. Therefore we get $\nu\left(F_{k}^{2}\right)=$ $2(m-1)>(m-1)+1=\nu\left(2 F_{k} F_{k+1}\right)$, which implies $\nu\left(F_{2 k}\right)=m$, establishing statement (a).

Moreover, since $F_{k+1}=1+2^{m-2}+a 2^{m-1}$ for some integer $a$, we get

$$
F_{2 k+1}=F_{k}^{2}+F_{k+1}^{2} \equiv 0+\left(1+2^{m-2}+a 2^{m-1}\right)^{2} \equiv 1+2^{m-1} \quad\left(\bmod 2^{m}\right)
$$

as desired in statement (c).
We are left to prove that $2^{m} \nmid F_{\ell}$ for $\ell<2 k$. Assume the contrary. Since $2^{m-1} \mid F_{\ell}$, from the induction hypothesis it follows that $\ell>k$. But then we have $F_{\ell}=F_{k-1} F_{\ell-k}+F_{k} F_{\ell-k+1}$, where the second summand is divisible by $2^{m-1}$ but the first one is not (since $F_{k-1}$ is odd and $\ell-k<k)$. Hence the sum is not divisible even by $2^{m-1}$. A contradiction.

Now, for every pair of integers $(a, b) \neq(0,0)$, let $\mu(a, b)=\min \{\nu(a), \nu(b)\}$. By an obvious induction, for every Fibonacci-type sequence $A=\left(a_{0}, a_{1}, \ldots\right)$ we have $\mu\left(a_{0}, a_{1}\right)=\mu\left(a_{1}, a_{2}\right)=\ldots$; denote this common value by $\mu(A)$. Also denote by $p_{n}(A)$ the period of this sequence modulo $2^{n}$, that is, the least $p>0$ such that $a_{k+p} \equiv a_{k}\left(\bmod 2^{n}\right)$ for all $k \geq 0$.
Lemma 3. Let $A=\left(a_{0}, a_{1}, \ldots\right)$ be a Fibonacci-type sequence such that $\mu(A)=k<n$. Then $p_{n}(A)=3 \cdot 2^{n-1-k}$.
Proof. First, we note that the sequence $\left(a_{0}, a_{1}, \ldots\right)$ has period $p$ modulo $2^{n}$ if and only if the sequence $\left(a_{0} / 2^{k}, a_{1} / 2^{k}, \ldots\right)$ has period $p$ modulo $2^{n-k}$. Hence, passing to this sequence we can assume that $k=0$.

We prove the statement by induction on $n$. It is easy to see that for $n=1,2$ the claim is true; actually, each Fibonacci-type sequence $A$ with $\mu(A)=0$ behaves as $0,1,1,0,1,1, \ldots$ modulo 2 , and as $0,1,1,2,3,1,0,1,1,2,3,1, \ldots$ modulo 4 (all pairs of residues from which at least one is odd appear as a pair of consecutive terms in this sequence).

Now suppose that $n \geq 3$ and consider an arbitrary Fibonacci-type sequence $A=\left(a_{0}, a_{1}, \ldots\right)$ with $\mu(A)=0$. Obviously we should have $p_{n-1}(A) \mid p_{n}(A)$, or, using the induction hypothesis, $s=3 \cdot 2^{n-2} \mid p_{n}(A)$. Next, we may suppose that $a_{0}$ is even; hence $a_{1}$ is odd, and $a_{0}=2 b_{0}$, $a_{1}=2 b_{1}+1$ for some integers $b_{0}, b_{1}$.

Consider the Fibonacci-type sequence $B=\left(b_{0}, b_{1}, \ldots\right)$ starting with $\left(b_{0}, b_{1}\right)$. Since $a_{0}=$ $2 b_{0}+F_{0}, a_{1}=2 b_{1}+F_{1}$, by an easy induction we get $a_{k}=2 b_{k}+F_{k}$ for all $k \geq 0$. By the induction hypothesis, we have $p_{n-1}(B) \mid s$, hence the sequence $\left(2 b_{0}, 2 b_{1}, \ldots\right)$ is $s$-periodic modulo $2^{n}$. On the other hand, by Lemma 2 we have $F_{s+1} \equiv 1+2^{n-1}\left(\bmod 2^{n}\right), F_{2 s} \equiv 0$ $\left(\bmod 2^{n}\right), F_{2 s+1} \equiv 1\left(\bmod 2^{n}\right)$, hence

$$
\begin{gathered}
a_{s+1}=2 b_{s+1}+F_{s+1} \equiv 2 b_{1}+1+2^{n-1} \not \equiv 2 b_{1}+1=a_{1} \quad\left(\bmod 2^{n}\right) \\
a_{2 s}=2 b_{2 s}+F_{2 s} \equiv 2 b_{0}+0=a_{0} \quad\left(\bmod 2^{n}\right) \\
a_{2 s+1}=2 b_{2 s+1}+F_{2 s+1} \equiv 2 b_{1}+1=a_{1} \quad\left(\bmod 2^{n}\right)
\end{gathered}
$$

The first line means that $A$ is not $s$-periodic, while the other two provide that $a_{2 s} \equiv a_{0}$, $a_{2 s+1} \equiv a_{1}$ and hence $a_{2 s+t} \equiv a_{t}$ for all $t \geq 0$. Hence $s\left|p_{n}(A)\right| 2 s$ and $p_{n}(A) \neq s$, which means that $p_{n}(A)=2 s$, as desired.

Finally, Lemma 3 provides a straightforward method of counting the number of cycles. Actually, take any number $0 \leq k \leq n-1$ and consider all the cells $(i, j)$ with $\mu(i, j)=k$. The total number of such cells is $2^{2(n-k)}-2^{2(n-k-1)}=3 \cdot 2^{2 n-2 k-2}$. On the other hand, they are split into cycles, and by Lemma 3 the length of each cycle is $3 \cdot 2^{n-1-k}$. Hence the number of cycles consisting of these cells is exactly $\frac{3 \cdot 2^{2 n-2 k-2}}{3 \cdot 2^{n-1-k}}=2^{n-k-1}$. Finally, there is only one cell $(0,0)$ which is not mentioned in the previous computation, and it forms a separate cycle. So the total number of cycles is

$$
1+\sum_{k=0}^{n-1} 2^{n-1-k}=1+\left(1+2+4+\cdots+2^{n-1}\right)=2^{n}
$$

Comment. We outline a different proof for the essential part of Lemma 3. That is, we assume that $k=0$ and show that in this case the period of $\left(a_{i}\right)$ modulo $2^{n}$ coincides with the period of the Fibonacci numbers modulo $2^{n}$; then the proof can be finished by the arguments from Lemma 2..

Note that $p$ is a (not necessarily minimal) period of the sequence $\left(a_{i}\right)$ modulo $2^{n}$ if and only if we have $a_{0} \equiv a_{p}\left(\bmod 2^{n}\right), a_{1} \equiv a_{p+1}\left(\bmod 2^{n}\right)$, that is,

$$
\begin{align*}
& a_{0} \equiv a_{p} \equiv F_{p-1} a_{0}+F_{p} a_{1}=F_{p}\left(a_{1}-a_{0}\right)+F_{p+1} a_{0} \quad\left(\bmod 2^{n}\right),  \tag{1}\\
& a_{1} \equiv a_{p+1}=F_{p} a_{0}+F_{p+1} a_{1} \quad\left(\bmod 2^{n}\right) .
\end{align*}
$$

Now, If $p$ is a period of $\left(F_{i}\right)$ then we have $F_{p} \equiv F_{0}=0\left(\bmod 2^{n}\right)$ and $F_{p+1} \equiv F_{1}=1\left(\bmod 2^{n}\right)$, which by (1) implies that $p$ is a period of $\left(a_{i}\right)$ as well.

Conversely, suppose that $p$ is a period of $\left(a_{i}\right)$. Combining the relations of (1) we get

$$
\begin{aligned}
0=a_{1} \cdot a_{0}-a_{0} \cdot a_{1} & \equiv a_{1}\left(F_{p}\left(a_{1}-a_{0}\right)+F_{p+1} a_{0}\right)-a_{0}\left(F_{p} a_{0}+F_{p+1} a_{1}\right) \\
& =F_{p}\left(a_{1}^{2}-a_{1} a_{0}-a_{0}^{2}\right) \quad\left(\bmod 2^{n}\right), \\
a_{1}^{2}-a_{1} a_{0}-a_{0}^{2}=\left(a_{1}-a_{0}\right) a_{1}-a_{0} \cdot a_{0} & \equiv\left(a_{1}-a_{0}\right)\left(F_{p} a_{0}+F_{p+1} a_{1}\right)-a_{0}\left(F_{p}\left(a_{1}-a_{0}\right)+F_{p+1} a_{0}\right) \\
& =F_{p+1}\left(a_{1}^{2}-a_{1} a_{0}-a_{0}^{2}\right) \quad\left(\bmod 2^{n}\right) .
\end{aligned}
$$

Since at least one of the numbers $a_{0}, a_{1}$ is odd, the number $a_{1}^{2}-a_{1} a_{0}-a_{0}^{2}$ is odd as well. Therefore the previous relations are equivalent with $F_{p} \equiv 0\left(\bmod 2^{n}\right)$ and $F_{p+1} \equiv 1\left(\bmod 2^{n}\right)$, which means exactly that $p$ is a period of $\left(F_{0}, F_{1}, \ldots\right)$ modulo $2^{n}$.

So, the sets of periods of $\left(a_{i}\right)$ and $\left(F_{i}\right)$ coincide, and hence the minimal periods coincide as well.

# $52^{\text {nd }}$ International Mathematical Olympiad 

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# 52nd International Mathematical Olympiad 12-24 July 2011 Amsterdam The Netherlands 

# Problem shortlist with solutions 

## IMO regulation: these shortlist problems have to be kept strictly confidential until IMO 2012.

## The problem selection committee

Bart de Smit (chairman), Ilya Bogdanov, Johan Bosman, Andries Brouwer, Gabriele Dalla Torre, Géza Kós, Hendrik Lenstra, Charles Leytem, Ronald van Luijk, Christian Reiher, Eckard Specht, Hans Sterk, Lenny Taelman

The committee gratefully acknowledges the receipt of 142 problem proposals by the following 46 countries:

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## Algebra

## A1

## A1

For any set $A=\left\{a_{1}, a_{2}, a_{3}, a_{4}\right\}$ of four distinct positive integers with sum $s_{A}=a_{1}+a_{2}+a_{3}+a_{4}$, let $p_{A}$ denote the number of pairs $(i, j)$ with $1 \leq i<j \leq 4$ for which $a_{i}+a_{j}$ divides $s_{A}$. Among all sets of four distinct positive integers, determine those sets $A$ for which $p_{A}$ is maximal.

## A2

## A2

Determine all sequences $\left(x_{1}, x_{2}, \ldots, x_{2011}\right)$ of positive integers such that for every positive integer $n$ there is an integer $a$ with

$$
x_{1}^{n}+2 x_{2}^{n}+\cdots+2011 x_{2011}^{n}=a^{n+1}+1 .
$$

## A3

Determine all pairs $(f, g)$ of functions from the set of real numbers to itself that satisfy

$$
g(f(x+y))=f(x)+(2 x+y) g(y)
$$

for all real numbers $x$ and $y$.

## A4

Determine all pairs $(f, g)$ of functions from the set of positive integers to itself that satisfy

$$
f^{g(n)+1}(n)+g^{f(n)}(n)=f(n+1)-g(n+1)+1
$$

for every positive integer $n$. Here, $f^{k}(n)$ means $\underbrace{f(f(\ldots f}_{k}(n) \ldots))$.

## A5

Prove that for every positive integer $n$, the set $\{2,3,4, \ldots, 3 n+1\}$ can be partitioned into $n$ triples in such a way that the numbers from each triple are the lengths of the sides of some obtuse triangle.

## A6

Let $f$ be a function from the set of real numbers to itself that satisfies

$$
f(x+y) \leq y f(x)+f(f(x))
$$

for all real numbers $x$ and $y$. Prove that $f(x)=0$ for all $x \leq 0$.

## A7

Let $a, b$, and $c$ be positive real numbers satisfying $\min (a+b, b+c, c+a)>\sqrt{2}$ and $a^{2}+b^{2}+c^{2}=3$. Prove that

$$
\frac{a}{(b+c-a)^{2}}+\frac{b}{(c+a-b)^{2}}+\frac{c}{(a+b-c)^{2}} \geq \frac{3}{(a b c)^{2}}
$$

## Combinatorics

## C1

Let $n>0$ be an integer. We are given a balance and $n$ weights of weight $2^{0}, 2^{1}, \ldots, 2^{n-1}$. In a sequence of $n$ moves we place all weights on the balance. In the first move we choose a weight and put it on the left pan. In each of the following moves we choose one of the remaining weights and we add it either to the left or to the right pan. Compute the number of ways in which we can perform these $n$ moves in such a way that the right pan is never heavier than the left pan.

## C2

Suppose that 1000 students are standing in a circle. Prove that there exists an integer $k$ with $100 \leq k \leq 300$ such that in this circle there exists a contiguous group of $2 k$ students, for which the first half contains the same number of girls as the second half.

## C3

Let $\mathcal{S}$ be a finite set of at least two points in the plane. Assume that no three points of $\mathcal{S}$ are collinear. By a windmill we mean a process as follows. Start with a line $\ell$ going through a point $P \in \mathcal{S}$. Rotate $\ell$ clockwise around the pivot $P$ until the line contains another point $Q$ of $\mathcal{S}$. The point $Q$ now takes over as the new pivot. This process continues indefinitely, with the pivot always being a point from $\mathcal{S}$.

Show that for a suitable $P \in \mathcal{S}$ and a suitable starting line $\ell$ containing $P$, the resulting windmill will visit each point of $\mathcal{S}$ as a pivot infinitely often.

## C4

Determine the greatest positive integer $k$ that satisfies the following property: The set of positive integers can be partitioned into $k$ subsets $A_{1}, A_{2}, \ldots, A_{k}$ such that for all integers $n \geq 15$ and all $i \in\{1,2, \ldots, k\}$ there exist two distinct elements of $A_{i}$ whose sum is $n$.

## C5

Let $m$ be a positive integer and consider a checkerboard consisting of $m$ by $m$ unit squares. At the midpoints of some of these unit squares there is an ant. At time 0 , each ant starts moving with speed 1 parallel to some edge of the checkerboard. When two ants moving in opposite directions meet, they both turn $90^{\circ}$ clockwise and continue moving with speed 1 . When more than two ants meet, or when two ants moving in perpendicular directions meet, the ants continue moving in the same direction as before they met. When an ant reaches one of the edges of the checkerboard, it falls off and will not re-appear.

Considering all possible starting positions, determine the latest possible moment at which the last ant falls off the checkerboard or prove that such a moment does not necessarily exist.

## C6

Let $n$ be a positive integer and let $W=\ldots x_{-1} x_{0} x_{1} x_{2} \ldots$ be an infinite periodic word consisting of the letters $a$ and $b$. Suppose that the minimal period $N$ of $W$ is greater than $2^{n}$.

A finite nonempty word $U$ is said to appear in $W$ if there exist indices $k \leq \ell$ such that $U=x_{k} x_{k+1} \ldots x_{\ell}$. A finite word $U$ is called ubiquitous if the four words $U a, U b, a U$, and $b U$ all appear in $W$. Prove that there are at least $n$ ubiquitous finite nonempty words.

## C7

On a square table of 2011 by 2011 cells we place a finite number of napkins that each cover a square of 52 by 52 cells. In each cell we write the number of napkins covering it, and we record the maximal number $k$ of cells that all contain the same nonzero number. Considering all possible napkin configurations, what is the largest value of $k$ ?

## Geometry

## G1

Let $A B C$ be an acute triangle. Let $\omega$ be a circle whose center $L$ lies on the side $B C$. Suppose that $\omega$ is tangent to $A B$ at $B^{\prime}$ and to $A C$ at $C^{\prime}$. Suppose also that the circumcenter $O$ of the triangle $A B C$ lies on the shorter arc $B^{\prime} C^{\prime}$ of $\omega$. Prove that the circumcircle of $A B C$ and $\omega$ meet at two points.

## G2

Let $A_{1} A_{2} A_{3} A_{4}$ be a non-cyclic quadrilateral. Let $O_{1}$ and $r_{1}$ be the circumcenter and the circumradius of the triangle $A_{2} A_{3} A_{4}$. Define $O_{2}, O_{3}, O_{4}$ and $r_{2}, r_{3}, r_{4}$ in a similar way. Prove that

$$
\frac{1}{O_{1} A_{1}^{2}-r_{1}^{2}}+\frac{1}{O_{2} A_{2}^{2}-r_{2}^{2}}+\frac{1}{O_{3} A_{3}^{2}-r_{3}^{2}}+\frac{1}{O_{4} A_{4}^{2}-r_{4}^{2}}=0 .
$$

## G3

Let $A B C D$ be a convex quadrilateral whose sides $A D$ and $B C$ are not parallel. Suppose that the circles with diameters $A B$ and $C D$ meet at points $E$ and $F$ inside the quadrilateral. Let $\omega_{E}$ be the circle through the feet of the perpendiculars from $E$ to the lines $A B, B C$, and $C D$. Let $\omega_{F}$ be the circle through the feet of the perpendiculars from $F$ to the lines $C D, D A$, and $A B$. Prove that the midpoint of the segment $E F$ lies on the line through the two intersection points of $\omega_{E}$ and $\omega_{F}$.

## G4

Let $A B C$ be an acute triangle with circumcircle $\Omega$. Let $B_{0}$ be the midpoint of $A C$ and let $C_{0}$ be the midpoint of $A B$. Let $D$ be the foot of the altitude from $A$, and let $G$ be the centroid of the triangle $A B C$. Let $\omega$ be a circle through $B_{0}$ and $C_{0}$ that is tangent to the circle $\Omega$ at a point $X \neq A$. Prove that the points $D, G$, and $X$ are collinear.

## G5

Let $A B C$ be a triangle with incenter $I$ and circumcircle $\omega$. Let $D$ and $E$ be the second intersection points of $\omega$ with the lines $A I$ and $B I$, respectively. The chord $D E$ meets $A C$ at a point $F$, and $B C$ at a point $G$. Let $P$ be the intersection point of the line through $F$ parallel to $A D$ and the line through $G$ parallel to $B E$. Suppose that the tangents to $\omega$ at $A$ and at $B$ meet at a point $K$. Prove that the three lines $A E, B D$, and $K P$ are either parallel or concurrent.

## G6

Let $A B C$ be a triangle with $A B=A C$, and let $D$ be the midpoint of $A C$. The angle bisector of $\angle B A C$ intersects the circle through $D, B$, and $C$ in a point $E$ inside the triangle $A B C$. The line $B D$ intersects the circle through $A, E$, and $B$ in two points $B$ and $F$. The lines $A F$ and $B E$ meet at a point $I$, and the lines $C I$ and $B D$ meet at a point $K$. Show that $I$ is the incenter of triangle $K A B$.

## G7

Let $A B C D E F$ be a convex hexagon all of whose sides are tangent to a circle $\omega$ with center $O$. Suppose that the circumcircle of triangle $A C E$ is concentric with $\omega$. Let $J$ be the foot of the perpendicular from $B$ to $C D$. Suppose that the perpendicular from $B$ to $D F$ intersects the line $E O$ at a point $K$. Let $L$ be the foot of the perpendicular from $K$ to $D E$. Prove that $D J=D L$.

## G8

Let $A B C$ be an acute triangle with circumcircle $\omega$. Let $t$ be a tangent line to $\omega$. Let $t_{a}$, $t_{b}$, and $t_{c}$ be the lines obtained by reflecting $t$ in the lines $B C, C A$, and $A B$, respectively. Show that the circumcircle of the triangle determined by the lines $t_{a}, t_{b}$, and $t_{c}$ is tangent to the circle $\omega$.

## Number Theory

## N1

For any integer $d>0$, let $f(d)$ be the smallest positive integer that has exactly $d$ positive divisors (so for example we have $f(1)=1, f(5)=16$, and $f(6)=12$ ). Prove that for every integer $k \geq 0$ the number $f\left(2^{k}\right)$ divides $f\left(2^{k+1}\right)$.

## N2

Consider a polynomial $P(x)=\left(x+d_{1}\right)\left(x+d_{2}\right) \cdot \ldots \cdot\left(x+d_{9}\right)$, where $d_{1}, d_{2}, \ldots, d_{9}$ are nine distinct integers. Prove that there exists an integer $N$ such that for all integers $x \geq N$ the number $P(x)$ is divisible by a prime number greater than 20 .

## N3

Let $n \geq 1$ be an odd integer. Determine all functions $f$ from the set of integers to itself such that for all integers $x$ and $y$ the difference $f(x)-f(y)$ divides $x^{n}-y^{n}$.

## N4

For each positive integer $k$, let $t(k)$ be the largest odd divisor of $k$. Determine all positive integers $a$ for which there exists a positive integer $n$ such that all the differences

$$
t(n+a)-t(n), \quad t(n+a+1)-t(n+1), \quad \ldots, \quad t(n+2 a-1)-t(n+a-1)
$$

are divisible by 4.

## N5

Let $f$ be a function from the set of integers to the set of positive integers. Suppose that for any two integers $m$ and $n$, the difference $f(m)-f(n)$ is divisible by $f(m-n)$. Prove that for all integers $m, n$ with $f(m) \leq f(n)$ the number $f(n)$ is divisible by $f(m)$.

## N6

Let $P(x)$ and $Q(x)$ be two polynomials with integer coefficients such that no nonconstant polynomial with rational coefficients divides both $P(x)$ and $Q(x)$. Suppose that for every positive integer $n$ the integers $P(n)$ and $Q(n)$ are positive, and $2^{Q(n)}-1$ divides $3^{P(n)}-1$. Prove that $Q(x)$ is a constant polynomial.

N7
Let $p$ be an odd prime number. For every integer $a$, define the number

$$
S_{a}=\frac{a}{1}+\frac{a^{2}}{2}+\cdots+\frac{a^{p-1}}{p-1} .
$$

Let $m$ and $n$ be integers such that

$$
S_{3}+S_{4}-3 S_{2}=\frac{m}{n}
$$

Prove that $p$ divides $m$.

## N8

Let $k$ be a positive integer and set $n=2^{k}+1$. Prove that $n$ is a prime number if and only if the following holds: there is a permutation $a_{1}, \ldots, a_{n-1}$ of the numbers $1,2, \ldots, n-1$ and a sequence of integers $g_{1}, g_{2}, \ldots, g_{n-1}$ such that $n$ divides $g_{i}^{a_{i}}-a_{i+1}$ for every $i \in\{1,2, \ldots, n-1\}$, where we set $a_{n}=a_{1}$.

## A1

For any set $A=\left\{a_{1}, a_{2}, a_{3}, a_{4}\right\}$ of four distinct positive integers with sum $s_{A}=a_{1}+a_{2}+a_{3}+a_{4}$, let $p_{A}$ denote the number of pairs $(i, j)$ with $1 \leq i<j \leq 4$ for which $a_{i}+a_{j}$ divides $s_{A}$. Among all sets of four distinct positive integers, determine those sets $A$ for which $p_{A}$ is maximal.

Answer. The sets $A$ for which $p_{A}$ is maximal are the sets the form $\{d, 5 d, 7 d, 11 d\}$ and $\{d, 11 d, 19 d, 29 d\}$, where $d$ is any positive integer. For all these sets $p_{A}$ is 4 .

Solution. Firstly, we will prove that the maximum value of $p_{A}$ is at most 4 . Without loss of generality, we may assume that $a_{1}<a_{2}<a_{3}<a_{4}$. We observe that for each pair of indices $(i, j)$ with $1 \leq i<j \leq 4$, the sum $a_{i}+a_{j}$ divides $s_{A}$ if and only if $a_{i}+a_{j}$ divides $s_{A}-\left(a_{i}+a_{j}\right)=a_{k}+a_{l}$, where $k$ and $l$ are the other two indices. Since there are 6 distinct pairs, we have to prove that at least two of them do not satisfy the previous condition. We claim that two such pairs are $\left(a_{2}, a_{4}\right)$ and $\left(a_{3}, a_{4}\right)$. Indeed, note that $a_{2}+a_{4}>a_{1}+a_{3}$ and $a_{3}+a_{4}>a_{1}+a_{2}$. Hence $a_{2}+a_{4}$ and $a_{3}+a_{4}$ do not divide $s_{A}$. This proves $p_{A} \leq 4$.

Now suppose $p_{A}=4$. By the previous argument we have

$$
\begin{array}{lll}
a_{1}+a_{4} \mid a_{2}+a_{3} & \text { and } & a_{2}+a_{3} \mid a_{1}+a_{4}, \\
a_{1}+a_{2} \mid a_{3}+a_{4} & \text { and } & a_{3}+a_{4} \nmid a_{1}+a_{2}, \\
a_{1}+a_{3} \mid a_{2}+a_{4} & \text { and } & a_{2}+a_{4} \nmid a_{1}+a_{3} .
\end{array}
$$

Hence, there exist positive integers $m$ and $n$ with $m>n \geq 2$ such that

$$
\left\{\begin{array}{l}
a_{1}+a_{4}=a_{2}+a_{3} \\
m\left(a_{1}+a_{2}\right)=a_{3}+a_{4} \\
n\left(a_{1}+a_{3}\right)=a_{2}+a_{4}
\end{array}\right.
$$

Adding up the first equation and the third one, we get $n\left(a_{1}+a_{3}\right)=2 a_{2}+a_{3}-a_{1}$. If $n \geq 3$, then $n\left(a_{1}+a_{3}\right)>3 a_{3}>2 a_{2}+a_{3}>2 a_{2}+a_{3}-a_{1}$. This is a contradiction. Therefore $n=2$. If we multiply by 2 the sum of the first equation and the third one, we obtain

$$
6 a_{1}+2 a_{3}=4 a_{2},
$$

while the sum of the first one and the second one is

$$
(m+1) a_{1}+(m-1) a_{2}=2 a_{3} .
$$

Adding up the last two equations we get

$$
(m+7) a_{1}=(5-m) a_{2} .
$$

It follows that $5-m \geq 1$, because the left-hand side of the last equation and $a_{2}$ are positive. Since we have $m>n=2$, the integer $m$ can be equal only to either 3 or 4 . Substituting $(3,2)$ and $(4,2)$ for $(m, n)$ and solving the previous system of equations, we find the families of solutions $\{d, 5 d, 7 d, 11 d\}$ and $\{d, 11 d, 19 d, 29 d\}$, where $d$ is any positive integer.

## A2

Determine all sequences $\left(x_{1}, x_{2}, \ldots, x_{2011}\right)$ of positive integers such that for every positive integer $n$ there is an integer $a$ with

$$
x_{1}^{n}+2 x_{2}^{n}+\cdots+2011 x_{2011}^{n}=a^{n+1}+1 .
$$

Answer. The only sequence that satisfies the condition is

$$
\left(x_{1}, \ldots, x_{2011}\right)=(1, k, \ldots, k) \quad \text { with } k=2+3+\cdots+2011=2023065 .
$$

Solution. Throughout this solution, the set of positive integers will be denoted by $\mathbb{Z}_{+}$.

Put $k=2+3+\cdots+2011=2023065$. We have

$$
1^{n}+2 k^{n}+\cdots 2011 k^{n}=1+k \cdot k^{n}=k^{n+1}+1
$$

for all $n$, so $(1, k, \ldots, k)$ is a valid sequence. We shall prove that it is the only one.
Let a valid sequence $\left(x_{1}, \ldots, x_{2011}\right)$ be given. For each $n \in \mathbb{Z}_{+}$we have some $y_{n} \in \mathbb{Z}_{+}$with

$$
x_{1}^{n}+2 x_{2}^{n}+\cdots+2011 x_{2011}^{n}=y_{n}^{n+1}+1 .
$$

Note that $x_{1}^{n}+2 x_{2}^{n}+\cdots+2011 x_{2011}^{n}<\left(x_{1}+2 x_{2}+\cdots+2011 x_{2011}\right)^{n+1}$, which implies that the sequence $\left(y_{n}\right)$ is bounded. In particular, there is some $y \in \mathbb{Z}_{+}$with $y_{n}=y$ for infinitely many $n$.

Let $m$ be the maximum of all the $x_{i}$. Grouping terms with equal $x_{i}$ together, the sum $x_{1}^{n}+$ $2 x_{2}^{n}+\cdots+2011 x_{2011}^{n}$ can be written as

$$
x_{1}^{n}+2 x_{2}^{n}+\cdots+x_{2011}^{n}=a_{m} m^{n}+a_{m-1}(m-1)^{n}+\cdots+a_{1}
$$

with $a_{i} \geq 0$ for all $i$ and $a_{1}+\cdots+a_{m}=1+2+\cdots+2011$. So there exist arbitrarily large values of $n$, for which

$$
\begin{equation*}
a_{m} m^{n}+\cdots+a_{1}-1-y \cdot y^{n}=0 . \tag{1}
\end{equation*}
$$

The following lemma will help us to determine the $a_{i}$ and $y$ :
Lemma. Let integers $b_{1}, \ldots, b_{N}$ be given and assume that there are arbitrarily large positive integers $n$ with $b_{1}+b_{2} 2^{n}+\cdots+b_{N} N^{n}=0$. Then $b_{i}=0$ for all $i$.

Proof. Suppose that not all $b_{i}$ are zero. We may assume without loss of generality that $b_{N} \neq 0$.

Dividing through by $N^{n}$ gives

$$
\left|b_{N}\right|=\left|b_{N-1}\left(\frac{N-1}{N}\right)^{n}+\cdots+b_{1}\left(\frac{1}{N}\right)^{n}\right| \leq\left(\left|b_{N-1}\right|+\cdots+\left|b_{1}\right|\right)\left(\frac{N-1}{N}\right)^{n}
$$

The expression $\left(\frac{N-1}{N}\right)^{n}$ can be made arbitrarily small for $n$ large enough, contradicting the assumption that $b_{N}$ be non-zero.

We obviously have $y>1$. Applying the lemma to (1) we see that $a_{m}=y=m, a_{1}=1$, and all the other $a_{i}$ are zero. This implies $\left(x_{1}, \ldots, x_{2011}\right)=(1, m, \ldots, m)$. But we also have $1+m=a_{1}+\cdots+a_{m}=1+\cdots+2011=1+k$ so $m=k$, which is what we wanted to show.

## A3

Determine all pairs $(f, g)$ of functions from the set of real numbers to itself that satisfy

$$
g(f(x+y))=f(x)+(2 x+y) g(y)
$$

for all real numbers $x$ and $y$.

Answer. Either both $f$ and $g$ vanish identically, or there exists a real number $C$ such that $f(x)=x^{2}+C$ and $g(x)=x$ for all real numbers $x$.

Solution. Clearly all these pairs of functions satisfy the functional equation in question, so it suffices to verify that there cannot be any further ones. Substituting $-2 x$ for $y$ in the given functional equation we obtain

$$
\begin{equation*}
g(f(-x))=f(x) \tag{1}
\end{equation*}
$$

Using this equation for $-x-y$ in place of $x$ we obtain

$$
\begin{equation*}
f(-x-y)=g(f(x+y))=f(x)+(2 x+y) g(y) . \tag{2}
\end{equation*}
$$

Now for any two real numbers $a$ and $b$, setting $x=-b$ and $y=a+b$ we get

$$
f(-a)=f(-b)+(a-b) g(a+b) .
$$

If $c$ denotes another arbitrary real number we have similarly

$$
f(-b)=f(-c)+(b-c) g(b+c)
$$

as well as

$$
f(-c)=f(-a)+(c-a) g(c+a) .
$$

Adding all these equations up, we obtain

$$
((a+c)-(b+c)) g(a+b)+((a+b)-(a+c)) g(b+c)+((b+c)-(a+b)) g(a+c)=0 .
$$

Now given any three real numbers $x, y$, and $z$ one may determine three reals $a, b$, and $c$ such that $x=b+c, y=c+a$, and $z=a+b$, so that we get

$$
(y-x) g(z)+(z-y) g(x)+(x-z) g(y)=0 .
$$

This implies that the three points $(x, g(x)),(y, g(y))$, and $(z, g(z))$ from the graph of $g$ are collinear. Hence that graph is a line, i.e., $g$ is either a constant or a linear function.

Let us write $g(x)=A x+B$, where $A$ and $B$ are two real numbers. Substituting $(0,-y)$ for $(x, y)$ in (21) and denoting $C=f(0)$, we have $f(y)=A y^{2}-B y+C$. Now, comparing the coefficients of $x^{2}$ in (11) we see that $A^{2}=A$, so $A=0$ or $A=1$.

If $A=0$, then (1) becomes $B=-B x+C$ and thus $B=C=0$, which provides the first of the two solutions mentioned above.

Now suppose $A=1$. Then (11) becomes $x^{2}-B x+C+B=x^{2}-B x+C$, so $B=0$. Thus, $g(x)=x$ and $f(x)=x^{2}+C$, which is the second solution from above.
Comment. Another way to show that $g(x)$ is either a constant or a linear function is the following. If we interchange $x$ and $y$ in the given functional equation and subtract this new equation from the given one, we obtain

$$
f(x)-f(y)=(2 y+x) g(x)-(2 x+y) g(y) .
$$

Substituting $(x, 0),(1, x)$, and $(0,1)$ for $(x, y)$, we get

$$
\begin{aligned}
& f(x)-f(0)=x g(x)-2 x g(0), \\
& f(1)-f(x)=(2 x+1) g(1)-(x+2) g(x), \\
& f(0)-f(1)=2 g(0)-g(1) .
\end{aligned}
$$

Taking the sum of these three equations and dividing by 2 , we obtain

$$
g(x)=x(g(1)-g(0))+g(0) .
$$

This proves that $g(x)$ is either a constant of a linear function.

## A4

Determine all pairs $(f, g)$ of functions from the set of positive integers to itself that satisfy

$$
f^{g(n)+1}(n)+g^{f(n)}(n)=f(n+1)-g(n+1)+1
$$

for every positive integer $n$. Here, $f^{k}(n)$ means $\underbrace{f(f(\ldots f}_{k}(n) \ldots))$.

Answer. The only pair $(f, g)$ of functions that satisfies the equation is given by $f(n)=n$ and $g(n)=1$ for all $n$.

Solution. The given relation implies

$$
\begin{equation*}
f\left(f^{g(n)}(n)\right)<f(n+1) \text { for all } n, \tag{1}
\end{equation*}
$$

which will turn out to be sufficient to determine $f$.
Let $y_{1}<y_{2}<\ldots$ be all the values attained by $f$ (this sequence might be either finite or infinite). We will prove that for every positive $n$ the function $f$ attains at least $n$ values, and we have (i) $)_{n}: f(x)=y_{n}$ if and only if $x=n$, and (ii) $)_{n}: y_{n}=n$. The proof will follow the scheme

$$
\begin{equation*}
(\mathrm{i})_{1},(\mathrm{ii})_{1},(\mathrm{i})_{2},(\mathrm{ii})_{2}, \ldots,(\mathrm{i})_{n},(\mathrm{ii})_{n}, \ldots \tag{2}
\end{equation*}
$$

To start, consider any $x$ such that $f(x)=y_{1}$. If $x>1$, then (1) reads $f\left(f^{g(x-1)}(x-1)\right)<y_{1}$, contradicting the minimality of $y_{1}$. So we have that $f(x)=y_{1}$ is equivalent to $x=1$, establishing $(\mathrm{i})_{1}$.

Next, assume that for some $n$ statement $(\mathrm{i})_{n}$ is established, as well as all the previous statements in (2). Note that these statements imply that for all $k \geq 1$ and $a<n$ we have $f^{k}(x)=a$ if and only if $x=a$.

Now, each value $y_{i}$ with $1 \leq i \leq n$ is attained at the unique integer $i$, so $y_{n+1}$ exists. Choose an arbitrary $x$ such that $f(x)=y_{n+1}$; we necessarily have $x>n$. Substituting $x-1$ into (1) we have $f\left(f^{g(x-1)}(x-1)\right)<y_{n+1}$, which implies

$$
\begin{equation*}
f^{g(x-1)}(x-1) \in\{1, \ldots, n\} \tag{3}
\end{equation*}
$$

Set $b=f^{g(x-1)}(x-1)$. If $b<n$ then we would have $x-1=b$ which contradicts $x>n$. So $b=n$, and hence $y_{n}=n$, which proves (ii) ${ }_{n}$. Next, from (i) ${ }_{n}$ we now get $f(k)=n \Longleftrightarrow k=n$, so removing all the iterations of $f$ in (3) we obtain $x-1=b=n$, which proves (i) $n_{n+1}$.

So, all the statements in (2) are valid and hence $f(n)=n$ for all $n$. The given relation between $f$ and $g$ now reads $n+g^{n}(n)=n+1-g(n+1)+1$ or $g^{n}(n)+g(n+1)=2$, from which it
immediately follows that we have $g(n)=1$ for all $n$.

Comment. Several variations of the above solution are possible. For instance, one may first prove by induction that the smallest $n$ values of $f$ are exactly $f(1)<\cdots<f(n)$ and proceed as follows. We certainly have $f(n) \geq n$ for all $n$. If there is an $n$ with $f(n)>n$, then $f(x)>x$ for all $x \geq n$. From this we conclude $f^{g(n)+1}(n)>f^{g(n)}(n)>\cdots>f(n)$. But we also have $f^{g(n)+1}<f(n+1)$. Having squeezed in a function value between $f(n)$ and $f(n+1)$, we arrive at a contradiction.

In any case, the inequality (1) plays an essential rôle.

## A5

Prove that for every positive integer $n$, the set $\{2,3,4, \ldots, 3 n+1\}$ can be partitioned into $n$ triples in such a way that the numbers from each triple are the lengths of the sides of some obtuse triangle.

Solution. Throughout the solution, we denote by $[a, b]$ the set $\{a, a+1, \ldots, b\}$. We say that $\{a, b, c\}$ is an obtuse triple if $a, b, c$ are the sides of some obtuse triangle.
We prove by induction on $n$ that there exists a partition of [2,3n+1] into $n$ obtuse triples $A_{i}$ $(2 \leq i \leq n+1)$ having the form $A_{i}=\left\{i, a_{i}, b_{i}\right\}$. For the base case $n=1$, one can simply set $A_{2}=\{2,3,4\}$. For the induction step, we need the following simple lemma.

Lemma. Suppose that the numbers $a<b<c$ form an obtuse triple, and let $x$ be any positive number. Then the triple $\{a, b+x, c+x\}$ is also obtuse.

Proof. The numbers $a<b+x<c+x$ are the sides of a triangle because $(c+x)-(b+x)=$ $c-b<a$. This triangle is obtuse since $(c+x)^{2}-(b+x)^{2}=(c-b)(c+b+2 x)>(c-b)(c+b)>a^{2}$.

Now we turn to the induction step. Let $n>1$ and put $t=\lfloor n / 2\rfloor<n$. By the induction hypothesis, there exists a partition of the set $[2,3 t+1]$ into $t$ obtuse triples $A_{i}^{\prime}=\left\{i, a_{i}^{\prime}, b_{i}^{\prime}\right\}$ $(i \in[2, t+1])$. For the same values of $i$, define $A_{i}=\left\{i, a_{i}^{\prime}+(n-t), b_{i}^{\prime}+(n-t)\right\}$. The constructed triples are obviously disjoint, and they are obtuse by the lemma. Moreover, we have

$$
\bigcup_{i=2}^{t+1} A_{i}=[2, t+1] \cup[n+2, n+2 t+1] .
$$

Next, for each $i \in[t+2, n+1]$, define $A_{i}=\{i, n+t+i, 2 n+i\}$. All these sets are disjoint, and

$$
\bigcup_{i=t+2}^{n+1} A_{i}=[t+2, n+1] \cup[n+2 t+2,2 n+t+1] \cup[2 n+t+2,3 n+1],
$$

so

$$
\bigcup_{i=2}^{n+1} A_{i}=[2,3 n+1]
$$

Thus, we are left to prove that the triple $A_{i}$ is obtuse for each $i \in[t+2, n+1]$.
Since $(2 n+i)-(n+t+i)=n-t<t+2 \leq i$, the elements of $A_{i}$ are the sides of a triangle. Next, we have
$(2 n+i)^{2}-(n+t+i)^{2}=(n-t)(3 n+t+2 i) \geq \frac{n}{2} \cdot(3 n+3(t+1)+1)>\frac{n}{2} \cdot \frac{9 n}{2} \geq(n+1)^{2} \geq i^{2}$,
so this triangle is obtuse. The proof is completed.

## A6

Let $f$ be a function from the set of real numbers to itself that satisfies

$$
\begin{equation*}
f(x+y) \leq y f(x)+f(f(x)) \tag{1}
\end{equation*}
$$

for all real numbers $x$ and $y$. Prove that $f(x)=0$ for all $x \leq 0$.

Solution 1. Substituting $y=t-x$, we rewrite (11) as

$$
\begin{equation*}
f(t) \leq t f(x)-x f(x)+f(f(x)) \tag{2}
\end{equation*}
$$

Consider now some real numbers $a, b$ and use (2) with $t=f(a), x=b$ as well as with $t=f(b)$, $x=a$. We get

$$
\begin{aligned}
& f(f(a))-f(f(b)) \leq f(a) f(b)-b f(b) \\
& f(f(b))-f(f(a)) \leq f(a) f(b)-a f(a)
\end{aligned}
$$

Adding these two inequalities yields

$$
2 f(a) f(b) \geq a f(a)+b f(b)
$$

Now, substitute $b=2 f(a)$ to obtain $2 f(a) f(b) \geq a f(a)+2 f(a) f(b)$, or $a f(a) \leq 0$. So, we get

$$
\begin{equation*}
f(a) \geq 0 \quad \text { for all } a<0 \tag{3}
\end{equation*}
$$

Now suppose $f(x)>0$ for some real number $x$. From (2) we immediately get that for every $t<\frac{x f(x)-f(f(x))}{f(x)}$ we have $f(t)<0$. This contradicts (3); therefore

$$
\begin{equation*}
f(x) \leq 0 \quad \text { for all real } x, \tag{4}
\end{equation*}
$$

and by (3) again we get $f(x)=0$ for all $x<0$.
We are left to find $f(0)$. Setting $t=x<0$ in (2) we get

$$
0 \leq 0-0+f(0)
$$

so $f(0) \geq 0$. Combining this with (4) we obtain $f(0)=0$.

Solution 2. We will also use the condition of the problem in form (2). For clarity we divide the argument into four steps.

Step 1. We begin by proving that $f$ attains nonpositive values only. Assume that there exist some real number $z$ with $f(z)>0$. Substituting $x=z$ into (2) and setting $A=f(z)$, $B=-z f(z)-f(f(z))$ we get $f(t) \leq A t+B$ for all real $t$. Hence, if for any positive real number $t$ we substitute $x=-t, y=t$ into (1), we get

$$
\begin{aligned}
f(0) & \leq t f(-t)+f(f(-t)) \leq t(-A t+B)+A f(-t)+B \\
& \leq-t(A t-B)+A(-A t+B)+B=-A t^{2}-\left(A^{2}-B\right) t+(A+1) B
\end{aligned}
$$

But surely this cannot be true if we take $t$ to be large enough. This contradiction proves that we have indeed $f(x) \leq 0$ for all real numbers $x$. Note that for this reason (1) entails

$$
\begin{equation*}
f(x+y) \leq y f(x) \tag{5}
\end{equation*}
$$

for all real numbers $x$ and $y$.
Step 2. We proceed by proving that $f$ has at least one zero. If $f(0)=0$, we are done. Otherwise, in view of Step 1 we get $f(0)<0$. Observe that (5) tells us now $f(y) \leq y f(0)$ for all real numbers $y$. Thus we can specify a positive real number $a$ that is so large that $f(a)^{2}>-f(0)$. Put $b=f(a)$ and substitute $x=b$ and $y=-b$ into (5); we learn $-b^{2}<f(0) \leq-b f(b)$, i.e. $b<f(b)$. Now we apply (2) to $x=b$ and $t=f(b)$, which yields

$$
f(f(b)) \leq(f(b)-b) f(b)+f(f(b))
$$

i.e. $f(b) \geq 0$. So in view of Step $1, b$ is a zero of $f$.

Step 3. Next we show that if $f(a)=0$ and $b<a$, then $f(b)=0$ as well. To see this, we just substitute $x=b$ and $y=a-b$ into (5), thus getting $f(b) \geq 0$, which suffices by Step 1 .

Step 4. By Step 3, the solution of the problem is reduced to showing $f(0)=0$. Pick any zero $r$ of $f$ and substitute $x=r$ and $y=-1$ into (1). Because of $f(r)=f(r-1)=0$ this gives $f(0) \geq 0$ and hence $f(0)=0$ by Step 1 again.

Comment 1. Both of these solutions also show $f(x) \leq 0$ for all real numbers $x$. As one can see from Solution 1, this task gets much easier if one already knows that $f$ takes nonnegative values for sufficiently small arguments. Another way of arriving at this statement, suggested by the proposer, is as follows:

Put $a=f(0)$ and substitute $x=0$ into (1). This gives $f(y) \leq a y+f(a)$ for all real numbers $y$. Thus if for any real number $x$ we plug $y=a-x$ into (11), we obtain

$$
f(a) \leq(a-x) f(x)+f(f(x)) \leq(a-x) f(x)+a f(x)+f(a)
$$

and hence $0 \leq(2 a-x) f(x)$. In particular, if $x<2 a$, then $f(x) \geq 0$.
Having reached this point, one may proceed almost exactly as in the first solution to deduce $f(x) \leq 0$ for all $x$. Afterwards the problem can be solved in a few lines as shown in steps 3 and 4 of the second
solution.
Comment 2. The original problem also contained the question whether a nonzero function satisfying the problem condition exists. Here we present a family of such functions.

Notice first that if $g:(0, \infty) \longrightarrow[0, \infty)$ denotes any function such that

$$
\begin{equation*}
g(x+y) \geq y g(x) \tag{6}
\end{equation*}
$$

for all positive real numbers $x$ and $y$, then the function $f$ given by

$$
f(x)= \begin{cases}-g(x) & \text { if } x>0  \tag{7}\\ 0 & \text { if } x \leq 0\end{cases}
$$

automatically satisfies (11). Indeed, we have $f(x) \leq 0$ and hence also $f(f(x))=0$ for all real numbers $x$. So (11) reduces to (5); moreover, this inequality is nontrivial only if $x$ and $y$ are positive. In this last case it is provided by (6).

Now it is not hard to come up with a nonzero function $g$ obeying (6). E.g. $g(z)=C e^{z}$ (where $C$ is a positive constant) fits since the inequality $e^{y}>y$ holds for all (positive) real numbers $y$. One may also consider the function $g(z)=e^{z}-1$; in this case, we even have that $f$ is continuous.

## A7

Let $a, b$, and $c$ be positive real numbers satisfying $\min (a+b, b+c, c+a)>\sqrt{2}$ and $a^{2}+b^{2}+c^{2}=3$. Prove that

$$
\begin{equation*}
\frac{a}{(b+c-a)^{2}}+\frac{b}{(c+a-b)^{2}}+\frac{c}{(a+b-c)^{2}} \geq \frac{3}{(a b c)^{2}} . \tag{1}
\end{equation*}
$$

Throughout both solutions, we denote the sums of the form $f(a, b, c)+f(b, c, a)+f(c, a, b)$ by $\sum f(a, b, c)$.

Solution 1. The condition $b+c>\sqrt{2}$ implies $b^{2}+c^{2}>1$, so $a^{2}=3-\left(b^{2}+c^{2}\right)<2$, i.e. $a<\sqrt{2}<b+c$. Hence we have $b+c-a>0$, and also $c+a-b>0$ and $a+b-c>0$ for similar reasons.
We will use the variant of HÖLDER's inequality

$$
\frac{x_{1}^{p+1}}{y_{1}^{p}}+\frac{x_{1}^{p+1}}{y_{1}^{p}}+\ldots+\frac{x_{n}^{p+1}}{y_{n}^{p}} \geq \frac{\left(x_{1}+x_{2}+\ldots+x_{n}\right)^{p+1}}{\left(y_{1}+y_{2}+\ldots+y_{n}\right)^{p}}
$$

which holds for all positive real numbers $p, x_{1}, x_{2}, \ldots, x_{n}, y_{1}, y_{2}, \ldots, y_{n}$. Applying it to the left-hand side of (11) with $p=2$ and $n=3$, we get

$$
\begin{equation*}
\sum \frac{a}{(b+c-a)^{2}}=\sum \frac{\left(a^{2}\right)^{3}}{a^{5}(b+c-a)^{2}} \geq \frac{\left(a^{2}+b^{2}+c^{2}\right)^{3}}{\left(\sum a^{5 / 2}(b+c-a)\right)^{2}}=\frac{27}{\left(\sum a^{5 / 2}(b+c-a)\right)^{2}} \tag{2}
\end{equation*}
$$

To estimate the denominator of the right-hand part, we use an instance of SchUR's inequality, namely

$$
\sum a^{3 / 2}(a-b)(a-c) \geq 0
$$

which can be rewritten as

$$
\sum a^{5 / 2}(b+c-a) \leq a b c(\sqrt{a}+\sqrt{b}+\sqrt{c})
$$

Moreover, by the inequality between the arithmetic mean and the fourth power mean we also have

$$
\left(\frac{\sqrt{a}+\sqrt{b}+\sqrt{c}}{3}\right)^{4} \leq \frac{a^{2}+b^{2}+c^{2}}{3}=1
$$

i.e., $\sqrt{a}+\sqrt{b}+\sqrt{c} \leq 3$. Hence, (2) yields

$$
\sum \frac{a}{(b+c-a)^{2}} \geq \frac{27}{(a b c(\sqrt{a}+\sqrt{b}+\sqrt{c}))^{2}} \geq \frac{3}{a^{2} b^{2} c^{2}}
$$

thus solving the problem.

Comment. In this solution, one may also start from the following version of HöLDER's inequality

$$
\left(\sum_{i=1}^{n} a_{i}^{3}\right)\left(\sum_{i=1}^{n} b_{i}^{3}\right)\left(\sum_{i=1}^{n} c_{i}^{3}\right) \geq\left(\sum_{i=1}^{n} a_{i} b_{i} c_{i}\right)^{3}
$$

applied as

$$
\sum \frac{a}{(b+c-a)^{2}} \cdot \sum a^{3}(b+c-a) \cdot \sum a^{2}(b+c-a) \geq 27 .
$$

After doing that, one only needs the slightly better known instances

$$
\sum a^{3}(b+c-a) \leq(a+b+c) a b c \quad \text { and } \quad \sum a^{2}(b+c-a) \leq 3 a b c
$$

of Schur's Inequality.

Solution 2. As in Solution 1, we mention that all the numbers $b+c-a, a+c-b, a+b-c$ are positive. We will use only this restriction and the condition

$$
\begin{equation*}
a^{5}+b^{5}+c^{5} \geq 3 \tag{3}
\end{equation*}
$$

which is weaker than the given one. Due to the symmetry we may assume that $a \geq b \geq c$. In view of (3), it suffices to prove the inequality

$$
\sum \frac{a^{3} b^{2} c^{2}}{(b+c-a)^{2}} \geq \sum a^{5}
$$

or, moving all the terms into the left-hand part,

$$
\begin{equation*}
\sum \frac{a^{3}}{(b+c-a)^{2}}\left((b c)^{2}-(a(b+c-a))^{2}\right) \geq 0 \tag{4}
\end{equation*}
$$

Note that the signs of the expressions $(y z)^{2}-(x(y+z-x))^{2}$ and $y z-x(y+z-x)=(x-y)(x-z)$ are the same for every positive $x, y, z$ satisfying the triangle inequality. So the terms in (4) corresponding to $a$ and $c$ are nonnegative, and hence it is sufficient to prove that the sum of the terms corresponding to $a$ and $b$ is nonnegative. Equivalently, we need the relation

$$
\frac{a^{3}}{(b+c-a)^{2}}(a-b)(a-c)(b c+a(b+c-a)) \geq \frac{b^{3}}{(a+c-b)^{2}}(a-b)(b-c)(a c+b(a+c-b)) .
$$

Obviously, we have

$$
a^{3} \geq b^{3} \geq 0, \quad 0<b+c-a \leq a+c-b, \quad \text { and } \quad a-c \geq b-c \geq 0,
$$

hence it suffices to prove that

$$
\frac{a b+a c+b c-a^{2}}{b+c-a} \geq \frac{a b+a c+b c-b^{2}}{c+a-b}
$$

Since all the denominators are positive, it is equivalent to

$$
(c+a-b)\left(a b+a c+b c-a^{2}\right)-\left(a b+a c+b c-b^{2}\right)(b+c-a) \geq 0
$$

or

$$
(a-b)\left(2 a b-a^{2}-b^{2}+a c+b c\right) \geq 0 .
$$

Since $a \geq b$, the last inequality follows from

$$
c(a+b)>(a-b)^{2}
$$

which holds since $c>a-b \geq 0$ and $a+b>a-b \geq 0$.

## C1

Let $n>0$ be an integer. We are given a balance and $n$ weights of weight $2^{0}, 2^{1}, \ldots, 2^{n-1}$. In a sequence of $n$ moves we place all weights on the balance. In the first move we choose a weight and put it on the left pan. In each of the following moves we choose one of the remaining weights and we add it either to the left or to the right pan. Compute the number of ways in which we can perform these $n$ moves in such a way that the right pan is never heavier than the left pan.

Answer. The number $f(n)$ of ways of placing the $n$ weights is equal to the product of all odd positive integers less than or equal to $2 n-1$, i.e. $f(n)=(2 n-1)!!=1 \cdot 3 \cdot 5 \cdot \ldots \cdot(2 n-1)$.

Solution 1. Assume $n \geq 2$. We claim

$$
\begin{equation*}
f(n)=(2 n-1) f(n-1) . \tag{1}
\end{equation*}
$$

Firstly, note that after the first move the left pan is always at least 1 heavier than the right one. Hence, any valid way of placing the $n$ weights on the scale gives rise, by not considering weight 1 , to a valid way of placing the weights $2,2^{2}, \ldots, 2^{n-1}$.
If we divide the weight of each weight by 2 , the answer does not change. So these $n-1$ weights can be placed on the scale in $f(n-1)$ valid ways. Now we look at weight 1 . If it is put on the scale in the first move, then it has to be placed on the left side, otherwise it can be placed either on the left or on the right side, because after the first move the difference between the weights on the left pan and the weights on the right pan is at least 2 . Hence, there are exactly $2 n-1$ different ways of inserting weight 1 in each of the $f(n-1)$ valid sequences for the $n-1$ weights in order to get a valid sequence for the $n$ weights. This proves the claim.

Since $f(1)=1$, by induction we obtain for all positive integers $n$

$$
f(n)=(2 n-1)!!=1 \cdot 3 \cdot 5 \cdot \ldots \cdot(2 n-1) .
$$

Comment 1. The word "compute" in the statement of the problem is probably too vague. An alternative but more artificial question might ask for the smallest $n$ for which the number of valid ways is divisible by 2011. In this case the answer would be 1006.

Comment 2. It is useful to remark that the answer is the same for any set of weights where each weight is heavier than the sum of the lighter ones. Indeed, in such cases the given condition is equivalent to asking that during the process the heaviest weight on the balance is always on the left pan.

Comment 3. Instead of considering the lightest weight, one may also consider the last weight put on the balance. If this weight is $2^{n-1}$ then it should be put on the left pan. Otherwise it may be put on
any pan; the inequality would not be violated since at this moment the heaviest weight is already put onto the left pan. In view of the previous comment, in each of these $2 n-1$ cases the number of ways to place the previous weights is exactly $f(n-1)$, which yields (1).

Solution 2. We present a different way of obtaining (11). Set $f(0)=1$. Firstly, we find a recurrent formula for $f(n)$.

Assume $n \geq 1$. Suppose that weight $2^{n-1}$ is placed on the balance in the $i$-th move with $1 \leq i \leq n$. This weight has to be put on the left pan. For the previous moves we have $\binom{n-1}{i-1}$ choices of the weights and from Comment 2 there are $f(i-1)$ valid ways of placing them on the balance. For later moves there is no restriction on the way in which the weights are to be put on the pans. Therefore, all $(n-i)!2^{n-i}$ ways are possible. This gives

$$
\begin{equation*}
f(n)=\sum_{i=1}^{n}\binom{n-1}{i-1} f(i-1)(n-i)!2^{n-i}=\sum_{i=1}^{n} \frac{(n-1)!f(i-1) 2^{n-i}}{(i-1)!} \tag{2}
\end{equation*}
$$

Now we are ready to prove (1). Using $n-1$ instead of $n$ in (2) we get

$$
f(n-1)=\sum_{i=1}^{n-1} \frac{(n-2)!f(i-1) 2^{n-1-i}}{(i-1)!}
$$

Hence, again from (2) we get

$$
\begin{aligned}
f(n)=2(n-1) \sum_{i=1}^{n-1} & \frac{(n-2)!f(i-1) 2^{n-1-i}}{(i-1)!}+f(n-1) \\
& =(2 n-2) f(n-1)+f(n-1)=(2 n-1) f(n-1)
\end{aligned}
$$

QED.

Comment. There exist different ways of obtaining the formula (2). Here we show one of them.
Suppose that in the first move we use weight $2^{n-i+1}$. Then the lighter $n-i$ weights may be put on the balance at any moment and on either pan. This gives $2^{n-i} \cdot(n-1)!/(i-1)!$ choices for the moves (moments and choices of pan) with the lighter weights. The remaining $i-1$ moves give a valid sequence for the $i-1$ heavier weights and this is the only requirement for these moves, so there are $f(i-1)$ such sequences. Summing over all $i=1,2, \ldots, n$ we again come to (22).

## C2

Suppose that 1000 students are standing in a circle. Prove that there exists an integer $k$ with $100 \leq k \leq 300$ such that in this circle there exists a contiguous group of $2 k$ students, for which the first half contains the same number of girls as the second half.

Solution. Number the students consecutively from 1 to 1000 . Let $a_{i}=1$ if the $i$ th student is a girl, and $a_{i}=0$ otherwise. We expand this notion for all integers $i$ by setting $a_{i+1000}=$ $a_{i-1000}=a_{i}$. Next, let

$$
S_{k}(i)=a_{i}+a_{i+1}+\cdots+a_{i+k-1} .
$$

Now the statement of the problem can be reformulated as follows:
There exist an integer $k$ with $100 \leq k \leq 300$ and an index $i$ such that $S_{k}(i)=S_{k}(i+k)$.
Assume now that this statement is false. Choose an index $i$ such that $S_{100}(i)$ attains the maximal possible value. In particular, we have $S_{100}(i-100)-S_{100}(i)<0$ and $S_{100}(i)-S_{100}(i+100)>0$, for if we had an equality, then the statement would hold. This means that the function $S(j)$ $S(j+100)$ changes sign somewhere on the segment $[i-100, i]$, so there exists some index $j \in$ [ $i-100, i-1]$ such that

$$
\begin{equation*}
S_{100}(j) \leq S_{100}(j+100)-1, \quad \text { but } \quad S_{100}(j+1) \geq S_{100}(j+101)+1 \tag{1}
\end{equation*}
$$

Subtracting the first inequality from the second one, we get $a_{j+100}-a_{j} \geq a_{j+200}-a_{j+100}+2$, so

$$
a_{j}=0, \quad a_{j+100}=1, \quad a_{j+200}=0
$$

Substituting this into the inequalities of (1), we also obtain $S_{99}(j+1) \leq S_{99}(j+101) \leq S_{99}(j+1)$, which implies

$$
\begin{equation*}
S_{99}(j+1)=S_{99}(j+101) \tag{2}
\end{equation*}
$$

Now let $k$ and $\ell$ be the least positive integers such that $a_{j-k}=1$ and $a_{j+200+\ell}=1$. By symmetry, we may assume that $k \geq \ell$. If $k \geq 200$ then we have $a_{j}=a_{j-1}=\cdots=a_{j-199}=0$, so $S_{100}(j-199)=S_{100}(j-99)=0$, which contradicts the initial assumption. Hence $\ell \leq k \leq 199$. Finally, we have

$$
\begin{gathered}
S_{100+\ell}(j-\ell+1)=\left(a_{j-\ell+1}+\cdots+a_{j}\right)+S_{99}(j+1)+a_{j+100}=S_{99}(j+1)+1 \\
S_{100+\ell}(j+101)=S_{99}(j+101)+\left(a_{j+200}+\cdots+a_{j+200+\ell-1}\right)+a_{j+200+\ell}=S_{99}(j+101)+1 .
\end{gathered}
$$

Comparing with (2) we get $S_{100+\ell}(j-\ell+1)=S_{100+\ell}(j+101)$ and $100+\ell \leq 299$, which again contradicts our assumption.

Comment. It may be seen from the solution that the number 300 from the problem statement can be
replaced by 299. Here we consider some improvements of this result. Namely, we investigate which interval can be put instead of $[100,300]$ in order to keep the problem statement valid.

First of all, the two examples

$$
\underbrace{1,1, \ldots, 1}_{167}, \underbrace{0,0, \ldots, 0}_{167}, \underbrace{1,1, \ldots, 1}_{167}, \underbrace{0,0, \ldots, 0}_{167}, \underbrace{1,1, \ldots, 1}_{167}, \underbrace{0,0, \ldots, 0}_{165}
$$

and

$$
\underbrace{1,1, \ldots, 1}_{249}, \underbrace{0,0, \ldots, 0}_{251}, \underbrace{1,1, \ldots, 1}_{249}, \underbrace{0,0, \ldots, 0}_{251}
$$

show that the interval can be changed neither to $[84,248]$ nor to $[126,374]$.
On the other hand, we claim that this interval can be changed to [125, 250]. Note that this statement is invariant under replacing all 1's by 0's and vice versa. Assume, to the contrary, that there is no admissible $k \in[125,250]$. The arguments from the solution easily yield the following lemma.

Lemma. Under our assumption, suppose that for some indices $i<j$ we have $S_{125}(i) \leq S_{125}(i+125)$ but $S_{125}(j) \geq S_{125}(j+125)$. Then there exists some $t \in[i, j-1]$ such that $a_{t}=a_{t-1}=\cdots=a_{t-125}=0$ and $a_{t+250}=a_{t+251}=\cdots=a_{t+375}=0$.

Let us call a segment $[i, j]$ of indices a crowd, if (a) $a_{i}=a_{i+1}=\cdots=a_{j}$, but $a_{i-1} \neq a_{i} \neq a_{j+1}$, and (b) $j-i \geq 125$. Now, using the lemma, one can get in the same way as in the solution that there exists some crowd. Take all the crowds in the circle, and enumerate them in cyclic order as $A_{1}, \ldots, A_{d}$. We also assume always that $A_{s+d}=A_{s-d}=A_{s}$.

Consider one of the crowds, say $A_{1}$. We have $A_{1}=[i, i+t]$ with $125 \leq t \leq 248$ (if $t \geq 249$, then $a_{i}=a_{i+1}=\cdots=a_{i+249}$ and therefore $S_{125}(i)=S_{125}(i+125)$, which contradicts our assumption). We may assume that $a_{i}=1$. Then we have $S_{125}(i+t-249) \leq 125=S_{125}(i+t-124)$ and $S_{125}(i)=125 \geq S_{125}(i+125)$, so by the lemma there exists some index $j \in[i+t-249, i-1]$ such that the segments $[j-125, j]$ and $[j+250, j+375]$ are contained in some crowds.

Let us fix such $j$ and denote the segment $[j+1, j+249]$ by $B_{1}$. Clearly, $A_{1} \subseteq B_{1}$. Moreover, $B_{1}$ cannot contain any crowd other than $A_{1}$ since $\left|B_{1}\right|=249<2 \cdot 126$. Hence it is clear that $j \in A_{d}$ and $j+250 \in A_{2}$. In particular, this means that the genders of $A_{d}$ and $A_{2}$ are different from that of $A_{1}$.
Performing this procedure for every crowd $A_{s}$, we find segments $B_{s}=\left[j_{s}+1, j_{s}+249\right]$ such that $\left|B_{s}\right|=249, A_{s} \subseteq B_{s}$, and $j_{s} \in A_{s-1}, j_{s}+250 \in A_{s+1}$. So, $B_{s}$ covers the whole segment between $A_{s-1}$ and $A_{s+1}$, hence the sets $B_{1}, \ldots, B_{d}$ cover some 1000 consecutive indices. This implies $249 d \geq 1000$, and $d \geq 5$. Moreover, the gender of $A_{i}$ is alternating, so $d$ is even; therefore $d \geq 6$.

Consider now three segments $A_{1}=\left[i_{1}, i_{1}^{\prime}\right], B_{2}=\left[j_{2}+1, j_{2}+249\right], A_{3}=\left[i_{3}, i_{3}^{\prime}\right]$. By construction, we have $\left[j_{2}-125, j_{2}\right] \subseteq A_{1}$ and $\left[j_{2}+250, j_{2}+375\right] \subseteq A_{3}$, whence $i_{1} \leq j_{2}-125, i_{3}^{\prime} \geq j_{2}+375$. Therefore $i_{3}^{\prime}-i_{1} \geq 500$. Analogously, if $A_{4}=\left[i_{4}, i_{4}^{\prime}\right], A_{6}=\left[i_{6}, i_{6}^{\prime}\right]$ then $i_{6}^{\prime}-i_{4} \geq 500$. But from $d \geq 6$ we get $i_{1}<i_{3}^{\prime}<i_{4}<i_{6}^{\prime}<i_{1}+1000$, so $1000>\left(i_{3}^{\prime}-i_{1}\right)+\left(i_{6}^{\prime}-i_{4}\right) \geq 500+500$. This final contradiction shows that our claim holds.

One may even show that the interval in the statement of the problem may be replaced by [125, 249] (both these numbers cannot be improved due to the examples above). But a proof of this fact is a bit messy, and we do not present it here.

## C3

Let $\mathcal{S}$ be a finite set of at least two points in the plane. Assume that no three points of $\mathcal{S}$ are collinear. By a windmill we mean a process as follows. Start with a line $\ell$ going through a point $P \in \mathcal{S}$. Rotate $\ell$ clockwise around the pivot $P$ until the line contains another point $Q$ of $\mathcal{S}$. The point $Q$ now takes over as the new pivot. This process continues indefinitely, with the pivot always being a point from $\mathcal{S}$.

Show that for a suitable $P \in \mathcal{S}$ and a suitable starting line $\ell$ containing $P$, the resulting windmill will visit each point of $\mathcal{S}$ as a pivot infinitely often.

Solution. Give the rotating line an orientation and distinguish its sides as the oranje side and the blue side. Notice that whenever the pivot changes from some point $T$ to another point $U$, after the change, $T$ is on the same side as $U$ was before. Therefore, the number of elements of $\mathcal{S}$ on the oranje side and the number of those on the blue side remain the same throughout the whole process (except for those moments when the line contains two points).


First consider the case that $|\mathcal{S}|=2 n+1$ is odd. We claim that through any point $T \in \mathcal{S}$, there is a line that has $n$ points on each side. To see this, choose an oriented line through $T$ containing no other point of $\mathcal{S}$ and suppose that it has $n+r$ points on its oranje side. If $r=0$ then we have established the claim, so we may assume that $r \neq 0$. As the line rotates through $180^{\circ}$ around $T$, the number of points of $\mathcal{S}$ on its oranje side changes by 1 whenever the line passes through a point; after $180^{\circ}$, the number of points on the oranje side is $n-r$. Therefore there is an intermediate stage at which the oranje side, and thus also the blue side, contains $n$ points.

Now select the point $P$ arbitrarily, and choose a line through $P$ that has $n$ points of $\mathcal{S}$ on each side to be the initial state of the windmill. We will show that during a rotation over $180^{\circ}$, the line of the windmill visits each point of $\mathcal{S}$ as a pivot. To see this, select any point $T$ of $\mathcal{S}$ and select a line $\ell$ through $T$ that separates $\mathcal{S}$ into equal halves. The point $T$ is the unique point of $\mathcal{S}$ through which a line in this direction can separate the points of $\mathcal{S}$ into equal halves (parallel translation would disturb the balance). Therefore, when the windmill line is parallel to $\ell$, it must be $\ell$ itself, and so pass through $T$.

Next suppose that $|\mathcal{S}|=2 n$. Similarly to the odd case, for every $T \in \mathcal{S}$ there is an oriented
line through $T$ with $n-1$ points on its oranje side and $n$ points on its blue side. Select such an oriented line through an arbitrary $P$ to be the initial state of the windmill.

We will now show that during a rotation over $360^{\circ}$, the line of the windmill visits each point of $\mathcal{S}$ as a pivot. To see this, select any point $T$ of $\mathcal{S}$ and an oriented line $\ell$ through $T$ that separates $\mathcal{S}$ into two subsets with $n-1$ points on its oranje and $n$ points on its blue side. Again, parallel translation would change the numbers of points on the two sides, so when the windmill line is parallel to $\ell$ with the same orientation, the windmill line must pass through $T$.

Comment. One may shorten this solution in the following way.
Suppose that $|\mathcal{S}|=2 n+1$. Consider any line $\ell$ that separates $\mathcal{S}$ into equal halves; this line is unique given its direction and contains some point $T \in \mathcal{S}$. Consider the windmill starting from this line. When the line has made a rotation of $180^{\circ}$, it returns to the same location but the oranje side becomes blue and vice versa. So, for each point there should have been a moment when it appeared as pivot, as this is the only way for a point to pass from on side to the other.

Now suppose that $|\mathcal{S}|=2 n$. Consider a line having $n-1$ and $n$ points on the two sides; it contains some point $T$. Consider the windmill starting from this line. After having made a rotation of $180^{\circ}$, the windmill line contains some different point $R$, and each point different from $T$ and $R$ has changed the color of its side. So, the windmill should have passed through all the points.

## C4

Determine the greatest positive integer $k$ that satisfies the following property: The set of positive integers can be partitioned into $k$ subsets $A_{1}, A_{2}, \ldots, A_{k}$ such that for all integers $n \geq 15$ and all $i \in\{1,2, \ldots, k\}$ there exist two distinct elements of $A_{i}$ whose sum is $n$.

Answer. The greatest such number $k$ is 3 .

Solution 1. There are various examples showing that $k=3$ does indeed have the property under consideration. E.g. one can take

$$
\begin{gathered}
A_{1}=\{1,2,3\} \cup\{3 m \mid m \geq 4\}, \\
A_{2}=\{4,5,6\} \cup\{3 m-1 \mid m \geq 4\}, \\
A_{3}=\{7,8,9\} \cup\{3 m-2 \mid m \geq 4\} .
\end{gathered}
$$

To check that this partition fits, we notice first that the sums of two distinct elements of $A_{i}$ obviously represent all numbers $n \geq 1+12=13$ for $i=1$, all numbers $n \geq 4+11=15$ for $i=2$, and all numbers $n \geq 7+10=17$ for $i=3$. So, we are left to find representations of the numbers 15 and 16 as sums of two distinct elements of $A_{3}$. These are $15=7+8$ and $16=7+9$.

Let us now suppose that for some $k \geq 4$ there exist sets $A_{1}, A_{2}, \ldots, A_{k}$ satisfying the given property. Obviously, the sets $A_{1}, A_{2}, A_{3}, A_{4} \cup \cdots \cup A_{k}$ also satisfy the same property, so one may assume $k=4$.

Put $B_{i}=A_{i} \cap\{1,2, \ldots, 23\}$ for $i=1,2,3,4$. Now for any index $i$ each of the ten numbers $15,16, \ldots, 24$ can be written as sum of two distinct elements of $B_{i}$. Therefore this set needs to contain at least five elements. As we also have $\left|B_{1}\right|+\left|B_{2}\right|+\left|B_{3}\right|+\left|B_{4}\right|=23$, there has to be some index $j$ for which $\left|B_{j}\right|=5$. Let $B_{j}=\left\{x_{1}, x_{2}, x_{3}, x_{4}, x_{5}\right\}$. Finally, now the sums of two distinct elements of $A_{j}$ representing the numbers $15,16, \ldots, 24$ should be exactly all the pairwise sums of the elements of $B_{j}$. Calculating the sum of these numbers in two different ways, we reach

$$
4\left(x_{1}+x_{2}+x_{3}+x_{4}+x_{5}\right)=15+16+\ldots+24=195 .
$$

Thus the number 195 should be divisible by 4, which is false. This contradiction completes our solution.

Comment. There are several variation of the proof that $k$ should not exceed 3. E.g., one may consider the sets $C_{i}=A_{i} \cap\{1,2, \ldots, 19\}$ for $i=1,2,3,4$. As in the previous solution one can show that for some index $j$ one has $\left|C_{j}\right|=4$, and the six pairwise sums of the elements of $C_{j}$ should represent all numbers $15,16, \ldots, 20$. Let $C_{j}=\left\{y_{1}, y_{2}, y_{3}, y_{4}\right\}$ with $y_{1}<y_{2}<y_{3}<y_{4}$. It is not hard to deduce
$C_{j}=\{7,8,9,11\}$, so in particular we have $1 \notin C_{j}$. Hence it is impossible to represent 21 as sum of two distinct elements of $A_{j}$, which completes our argument.

Solution 2. Again we only prove that $k \leq 3$. Assume that $A_{1}, A_{2}, \ldots, A_{k}$ is a partition satisfying the given property. We construct a graph $\mathcal{G}$ on the set $V=\{1,2, \ldots, 18\}$ of vertices as follows. For each $i \in\{1,2, \ldots, k\}$ and each $d \in\{15,16,17,19\}$ we choose one pair of distinct elements $a, b \in A_{i}$ with $a+b=d$, and we draw an edge in the $i^{\text {th }}$ color connecting $a$ with $b$. By hypothesis, $\mathcal{G}$ has exactly 4 edges of each color.

Claim. The graph $\mathcal{G}$ contains at most one circuit.
Proof. Note that all the connected components of $\mathcal{G}$ are monochromatic and hence contain at most four edges. Thus also all circuits of $\mathcal{G}$ are monochromatic and have length at most four. Moreover, each component contains at most one circuit since otherwise it should contain at least five edges.

Suppose that there is a 4 -cycle in $\mathcal{G}$, say with vertices $a, b, c$, and $d$ in order. Then $\{a+b, b+$ $c, c+d, d+a\}=\{15,16,17,19\}$. Taking sums we get $2(a+b+c+d)=15+16+17+19$ which is impossible for parity reasons. Thus all circuits of $\mathcal{G}$ are triangles.

Now if the vertices $a, b$, and $c$ form such a triangle, then by a similar reasoning the set $\{a+b, b+$ $c, c+a\}$ coincides with either $\{15,16,17\}$, or $\{15,16,19\}$, or $\{16,17,19\}$, or $\{15,17,19\}$. The last of these alternatives can be excluded for parity reasons again, whilst in the first three cases the set $\{a, b, c\}$ appears to be either $\{7,8,9\}$, or $\{6,9,10\}$, or $\{7,9,10\}$, respectively. Thus, a component containing a circuit should contain 9 as a vertex. Therefore there is at most one such component and hence at most one circuit.

By now we know that $\mathcal{G}$ is a graph with $4 k$ edges, at least $k$ components and at most one circuit. Consequently, $\mathcal{G}$ must have at least $4 k+k-1$ vertices. Thus $5 k-1 \leq 18$, and $k \leq 3$.

## C5

Let $m$ be a positive integer and consider a checkerboard consisting of $m$ by $m$ unit squares. At the midpoints of some of these unit squares there is an ant. At time 0 , each ant starts moving with speed 1 parallel to some edge of the checkerboard. When two ants moving in opposite directions meet, they both turn $90^{\circ}$ clockwise and continue moving with speed 1 . When more than two ants meet, or when two ants moving in perpendicular directions meet, the ants continue moving in the same direction as before they met. When an ant reaches one of the edges of the checkerboard, it falls off and will not re-appear.

Considering all possible starting positions, determine the latest possible moment at which the last ant falls off the checkerboard or prove that such a moment does not necessarily exist.

Antswer. The latest possible moment for the last ant to fall off is $\frac{3 m}{2}-1$.

Solution. For $m=1$ the answer is clearly correct, so assume $m>1$. In the sequel, the word collision will be used to denote meeting of exactly two ants, moving in opposite directions.

If at the beginning we place an ant on the southwest corner square facing east and an ant on the southeast corner square facing west, then they will meet in the middle of the bottom row at time $\frac{m-1}{2}$. After the collision, the ant that moves to the north will stay on the board for another $m-\frac{1}{2}$ time units and thus we have established an example in which the last ant falls off at time $\frac{m-1}{2}+m-\frac{1}{2}=\frac{3 m}{2}-1$. So, we are left to prove that this is the latest possible moment.

Consider any collision of two ants $a$ and $a^{\prime}$. Let us change the rule for this collision, and enforce these two ants to turn anticlockwise. Then the succeeding behavior of all the ants does not change; the only difference is that $a$ and $a^{\prime}$ swap their positions. These arguments may be applied to any collision separately, so we may assume that at any collision, either both ants rotate clockwise or both of them rotate anticlockwise by our own choice.

For instance, we may assume that there are only two types of ants, depending on their initial direction: NE-ants, which move only north or east, and $S W$-ants, moving only south and west. Then we immediately obtain that all ants will have fallen off the board after $2 m-1$ time units. However, we can get a better bound by considering the last moment at which a given ant collides with another ant.

Choose a coordinate system such that the corners of the checkerboard are $(0,0),(m, 0),(m, m)$ and $(0, m)$. At time $t$, there will be no NE-ants in the region $\{(x, y): x+y<t+1\}$ and no SW-ants in the region $\{(x, y): x+y>2 m-t-1\}$. So if two ants collide at $(x, y)$ at time $t$, we have

$$
\begin{equation*}
t+1 \leq x+y \leq 2 m-t-1 \tag{1}
\end{equation*}
$$

Analogously, we may change the rules so that each ant would move either alternatingly north and west, or alternatingly south and east. By doing so, we find that apart from (1) we also have $|x-y| \leq m-t-1$ for each collision at point $(x, y)$ and time $t$.

To visualize this, put

$$
B(t)=\left\{(x, y) \in[0, m]^{2}: t+1 \leq x+y \leq 2 m-t-1 \text { and }|x-y| \leq m-t-1\right\} .
$$

An ant can thus only collide with another ant at time $t$ if it happens to be in the region $B(t)$. The following figure displays $B(t)$ for $t=\frac{1}{2}$ and $t=\frac{7}{2}$ in the case $m=6$ :


Now suppose that an NE-ant has its last collision at time $t$ and that it does so at the point ( $x, y$ ) (if the ant does not collide at all, it will fall off the board within $m-\frac{1}{2}<\frac{3 m}{2}-1$ time units, so this case can be ignored). Then we have $(x, y) \in B(t)$ and thus $x+y \geq t+1$ and $x-y \geq-(m-t-1)$. So we get

$$
x \geq \frac{(t+1)-(m-t-1)}{2}=t+1-\frac{m}{2} .
$$

By symmetry we also have $y \geq t+1-\frac{m}{2}$, and hence $\min \{x, y\} \geq t+1-\frac{m}{2}$. After this collision, the ant will move directly to an edge, which will take at most $m-\min \{x, y\}$ units of time. In sum, the total amount of time the ant stays on the board is at most

$$
t+(m-\min \{x, y\}) \leq t+m-\left(t+1-\frac{m}{2}\right)=\frac{3 m}{2}-1
$$

By symmetry, the same bound holds for SW-ants as well.

## C6

Let $n$ be a positive integer and let $W=\ldots x_{-1} x_{0} x_{1} x_{2} \ldots$ be an infinite periodic word consisting of the letters $a$ and $b$. Suppose that the minimal period $N$ of $W$ is greater than $2^{n}$.

A finite nonempty word $U$ is said to appear in $W$ if there exist indices $k \leq \ell$ such that $U=x_{k} x_{k+1} \ldots x_{\ell}$. A finite word $U$ is called ubiquitous if the four words $U a, U b, a U$, and $b U$ all appear in $W$. Prove that there are at least $n$ ubiquitous finite nonempty words.

Solution. Throughout the solution, all the words are nonempty. For any word $R$ of length $m$, we call the number of indices $i \in\{1,2, \ldots, N\}$ for which $R$ coincides with the subword $x_{i+1} x_{i+2} \ldots x_{i+m}$ of $W$ the multiplicity of $R$ and denote it by $\mu(R)$. Thus a word $R$ appears in $W$ if and only if $\mu(R)>0$. Since each occurrence of a word in $W$ is both succeeded by either the letter $a$ or the letter $b$ and similarly preceded by one of those two letters, we have

$$
\begin{equation*}
\mu(R)=\mu(R a)+\mu(R b)=\mu(a R)+\mu(b R) \tag{1}
\end{equation*}
$$

for all words $R$.
We claim that the condition that $N$ is in fact the minimal period of $W$ guarantees that each word of length $N$ has multiplicity 1 or 0 depending on whether it appears or not. Indeed, if the words $x_{i+1} x_{i+2} \ldots x_{i+N}$ and $x_{j+1} \ldots x_{j+N}$ are equal for some $1 \leq i<j \leq N$, then we have $x_{i+a}=x_{j+a}$ for every integer $a$, and hence $j-i$ is also a period.

Moreover, since $N>2^{n}$, at least one of the two words $a$ and $b$ has a multiplicity that is strictly larger than $2^{n-1}$.

For each $k=0,1, \ldots, n-1$, let $U_{k}$ be a subword of $W$ whose multiplicity is strictly larger than $2^{k}$ and whose length is maximal subject to this property. Note that such a word exists in view of the two observations made in the two previous paragraphs.

Fix some index $k \in\{0,1, \ldots, n-1\}$. Since the word $U_{k} b$ is longer than $U_{k}$, its multiplicity can be at most $2^{k}$, so in particular $\mu\left(U_{k} b\right)<\mu\left(U_{k}\right)$. Therefore, the word $U_{k} a$ has to appear by (11). For a similar reason, the words $U_{k} b, a U_{k}$, and $b U_{k}$ have to appear as well. Hence, the word $U_{k}$ is ubiquitous. Moreover, if the multiplicity of $U_{k}$ were strictly greater than $2^{k+1}$, then by (1) at least one of the two words $U_{k} a$ and $U_{k} b$ would have multiplicity greater than $2^{k}$ and would thus violate the maximality condition imposed on $U_{k}$.

So we have $\mu\left(U_{0}\right) \leq 2<\mu\left(U_{1}\right) \leq 4<\ldots \leq 2^{n-1}<\mu\left(U_{n-1}\right)$, which implies in particular that the words $U_{0}, U_{1}, \ldots, U_{n-1}$ have to be distinct. As they have been proved to be ubiquitous as well, the problem is solved.

Comment 1. There is an easy construction for obtaining ubiquitous words from appearing words whose multiplicity is at least two. Starting with any such word $U$ we may simply extend one of its occurrences in $W$ forwards and backwards as long as its multiplicity remains fixed, thus arriving at a
word that one might call the ubiquitous prolongation $p(U)$ of $U$.
There are several variants of the argument in the second half of the solution using the concept of prolongation. For instance, one may just take all ubiquitous words $U_{1}, U_{2}, \ldots, U_{\ell}$ ordered by increasing multiplicity and then prove for $i \in\{1,2, \ldots, \ell\}$ that $\mu\left(U_{i}\right) \leq 2^{i}$. Indeed, assume that $i$ is a minimal counterexample to this statement; then by the arguments similar to those presented above, the ubiquitous prolongation of one of the words $U_{i} a, U_{i} b, a U_{i}$ or $b U_{i}$ violates the definition of $U_{i}$.

Now the multiplicity of one of the two letters $a$ and $b$ is strictly greater than $2^{n-1}$, so passing to ubiquitous prolongations once more we obtain $2^{n-1}<\mu\left(U_{\ell}\right) \leq 2^{\ell}$, which entails $\ell \geq n$, as needed.

Comment 2. The bound $n$ for the number of ubiquitous subwords in the problem statement is not optimal, but it is close to an optimal one in the following sense. There is a universal constant $C>0$ such that for each positive integer $n$ there exists an infinite periodic word $W$ whose minimal period is greater than $2^{n}$ but for which there exist fewer than $C n$ ubiquitous words.

## C7

On a square table of 2011 by 2011 cells we place a finite number of napkins that each cover a square of 52 by 52 cells. In each cell we write the number of napkins covering it, and we record the maximal number $k$ of cells that all contain the same nonzero number. Considering all possible napkin configurations, what is the largest value of $k$ ?

Answer. $2011^{2}-\left(\left(52^{2}-35^{2}\right) \cdot 39-17^{2}\right)=4044121-57392=3986729$.

Solution 1. Let $m=39$, then $2011=52 m-17$. We begin with an example showing that there can exist 3986729 cells carrying the same positive number.


To describe it, we number the columns from the left to the right and the rows from the bottom to the top by $1,2, \ldots, 2011$. We will denote each napkin by the coordinates of its lowerleft cell. There are four kinds of napkins: first, we take all napkins $(52 i+36,52 j+1)$ with $0 \leq j \leq i \leq m-2$; second, we use all napkins $(52 i+1,52 j+36)$ with $0 \leq i \leq j \leq m-2$; third, we use all napkins $(52 i+36,52 i+36)$ with $0 \leq i \leq m-2$; and finally the napkin $(1,1)$. Different groups of napkins are shown by different types of hatchings in the picture.

Now except for those squares that carry two or more different hatchings, all squares have the number 1 written into them. The number of these exceptional cells is easily computed to be $\left(52^{2}-35^{2}\right) m-17^{2}=57392$.

We are left to prove that 3986729 is an upper bound for the number of cells containing the same number. Consider any configuration of napkins and any positive integer $M$. Suppose there are $g$ cells with a number different from $M$. Then it suffices to show $g \geq 57392$. Throughout the solution, a line will mean either a row or a column.

Consider any line $\ell$. Let $a_{1}, \ldots, a_{52 m-17}$ be the numbers written into its consecutive cells. For $i=1,2, \ldots, 52$, let $s_{i}=\sum_{t \equiv i(\bmod 52)} a_{t}$. Note that $s_{1}, \ldots, s_{35}$ have $m$ terms each, while $s_{36}, \ldots, s_{52}$ have $m-1$ terms each. Every napkin intersecting $\ell$ contributes exactly 1 to each $s_{i}$;
hence the number $s$ of all those napkins satisfies $s_{1}=\cdots=s_{52}=s$. Call the line $\ell$ rich if $s>(m-1) M$ and poor otherwise.
Suppose now that $\ell$ is rich. Then in each of the sums $s_{36}, \ldots, s_{52}$ there exists a term greater than $M$; consider all these terms and call the corresponding cells the rich bad cells for this line. So, each rich line contains at least 17 cells that are bad for this line.

If, on the other hand, $\ell$ is poor, then certainly $s<m M$ so in each of the sums $s_{1}, \ldots, s_{35}$ there exists a term less than $M$; consider all these terms and call the corresponding cells the poor bad cells for this line. So, each poor line contains at least 35 cells that are bad for this line.

Let us call all indices congruent to $1,2, \ldots$, or 35 modulo 52 small, and all other indices, i.e. those congruent to $36,37, \ldots$, or 52 modulo 52 , big. Recall that we have numbered the columns from the left to the right and the rows from the bottom to the top using the numbers $1,2, \ldots, 52 m-17$; we say that a line is big or small depending on whether its index is big or small. By definition, all rich bad cells for the rows belong to the big columns, while the poor ones belong to the small columns, and vice versa.

In each line, we put a strawberry on each cell that is bad for this line. In addition, for each small rich line we put an extra strawberry on each of its (rich) bad cells. A cell gets the strawberries from its row and its column independently.

Notice now that a cell with a strawberry on it contains a number different from $M$. If this cell gets a strawberry by the extra rule, then it contains a number greater than $M$. Moreover, it is either in a small row and in a big column, or vice versa. Suppose that it is in a small row, then it is not bad for its column. So it has not more than two strawberries in this case. On the other hand, if the extra rule is not applied to some cell, then it also has not more than two strawberries. So, the total number $N$ of strawberries is at most $2 g$.

We shall now estimate $N$ in a different way. For each of the $2 \cdot 35 \mathrm{~m}$ small lines, we have introduced at least 34 strawberries if it is rich and at least 35 strawberries if it is poor, so at least 34 strawberries in any case. Similarly, for each of the $2 \cdot 17(m-1)$ big lines, we put at least $\min (17,35)=17$ strawberries. Summing over all lines we obtain

$$
2 g \geq N \geq 2(35 m \cdot 34+17(m-1) \cdot 17)=2(1479 m-289)=2 \cdot 57392
$$

as desired.

Comment. The same reasoning applies also if we replace 52 by $R$ and 2011 by $R m-H$, where $m, R$, and $H$ are integers with $m, R \geq 1$ and $0 \leq H \leq \frac{1}{3} R$. More detailed information is provided after the next solution.

Solution 2. We present a different proof of the estimate which is the hard part of the problem. Let $S=35, H=17, m=39$; so the table size is $2011=S m+H(m-1)$, and the napkin size is $52=S+H$. Fix any positive integer $M$ and call a cell vicious if it contains a number distinct
from $M$. We will prove that there are at least $H^{2}(m-1)+2 S H m$ vicious cells.
Firstly, we introduce some terminology. As in the previous solution, we number rows and columns and we use the same notions of small and big indices and lines; so, an index is small if it is congruent to one of the numbers $1,2, \ldots, S$ modulo $(S+H)$. The numbers $1,2, \ldots, S+H$ will be known as residues. For two residues $i$ and $j$, we say that a cell is of type $(i, j)$ if the index of its row is congruent to $i$ and the index of its column to $j$ modulo $(S+H)$. The number of vicious cells of this type is denoted by $v_{i j}$.

Let $s, s^{\prime}$ be two variables ranging over small residues and let $h, h^{\prime}$ be two variables ranging over big residues. A cell is said to be of class $A, B, C$, or $D$ if its type is of shape $\left(s, s^{\prime}\right),(s, h),(h, s)$, or $\left(h, h^{\prime}\right)$, respectively. The numbers of vicious cells belonging to these classes are denoted in this order by $a, b, c$, and $d$. Observe that each cell belongs to exactly one class.

Claim 1. We have

$$
\begin{equation*}
m \leq \frac{a}{S^{2}}+\frac{b+c}{2 S H} . \tag{1}
\end{equation*}
$$

Proof. Consider an arbitrary small row $r$. Denote the numbers of vicious cells on $r$ belonging to the classes $A$ and $B$ by $\alpha$ and $\beta$, respectively. As in the previous solution, we obtain that $\alpha \geq S$ or $\beta \geq H$. So in each case we have $\frac{\alpha}{S}+\frac{\beta}{H} \geq 1$.

Performing this argument separately for each small row and adding up all the obtained inequalities, we get $\frac{a}{S}+\frac{b}{H} \geq m S$. Interchanging rows and columns we similarly get $\frac{a}{S}+\frac{c}{H} \geq m S$. Summing these inequalities and dividing by $2 S$ we get what we have claimed.

Claim 2. Fix two small residue $s, s^{\prime}$ and two big residues $h, h^{\prime}$. Then $2 m-1 \leq v_{s s^{\prime}}+v_{s h^{\prime}}+v_{h h^{\prime}}$. Proof. Each napkin covers exactly one cell of type ( $s, s^{\prime}$ ). Removing all napkins covering a vicious cell of this type, we get another collection of napkins, which covers each cell of type $\left(s, s^{\prime}\right)$ either 0 or $M$ times depending on whether the cell is vicious or not. Hence $\left(m^{2}-v_{s s^{\prime}}\right) M$ napkins are left and throughout the proof of Claim 2 we will consider only these remaining napkins. Now, using a red pen, write in each cell the number of napkins covering it. Notice that a cell containing a red number greater than $M$ is surely vicious.

We call two cells neighbors if they can be simultaneously covered by some napkin. So, each cell of type $\left(h, h^{\prime}\right)$ has not more than four neighbors of type $\left(s, s^{\prime}\right)$, while each cell of type $\left(s, h^{\prime}\right)$ has not more than two neighbors of each of the types $\left(s, s^{\prime}\right)$ and $\left(h, h^{\prime}\right)$. Therefore, each red number at a cell of type ( $h, h^{\prime}$ ) does not exceed $4 M$, while each red number at a cell of type $\left(s, h^{\prime}\right)$ does not exceed $2 M$.

Let $x, y$, and $z$ be the numbers of cells of type ( $h, h^{\prime}$ ) whose red number belongs to ( $M, 2 M$ ], $(2 M, 3 M]$, and $(3 M, 4 M]$, respectively. All these cells are vicious, hence $x+y+z \leq v_{h h^{\prime}}$. The red numbers appearing in cells of type $\left(h, h^{\prime}\right)$ clearly sum up to $\left(m^{2}-v_{s s^{\prime}}\right) M$. Bounding each of these numbers by a multiple of $M$ we get

$$
\left(m^{2}-v_{s s^{\prime}}\right) M \leq\left((m-1)^{2}-(x+y+z)\right) M+2 x M+3 y M+4 z M
$$

i.e.

$$
2 m-1 \leq v_{s s^{\prime}}+x+2 y+3 z \leq v_{s s^{\prime}}+v_{h h^{\prime}}+y+2 z
$$

So, to prove the claim it suffices to prove that $y+2 z \leq v_{s h^{\prime}}$.
For a cell $\delta$ of type $\left(h, h^{\prime}\right)$ and a cell $\beta$ of type $\left(s, h^{\prime}\right)$ we say that $\delta$ forces $\beta$ if there are more than $M$ napkins covering both of them. Since each red number in a cell of type $\left(s, h^{\prime}\right)$ does not exceed $2 M$, it cannot be forced by more than one cell.

On the other hand, if a red number in a $\left(h, h^{\prime}\right)$-cell belongs to $(2 M, 3 M]$, then it forces at least one of its neighbors of type $\left(s, h^{\prime}\right)$ (since the sum of red numbers in their cells is greater than $2 M)$. Analogously, an $\left(h, h^{\prime}\right)$-cell with the red number in $(3 M, 4 M]$ forces both its neighbors of type $\left(s, h^{\prime}\right)$, since their red numbers do not exceed $2 M$. Therefore there are at least $y+2 z$ forced cells and clearly all of them are vicious, as desired.

Claim 3. We have

$$
\begin{equation*}
2 m-1 \leq \frac{a}{S^{2}}+\frac{b+c}{2 S H}+\frac{d}{H^{2}} \tag{2}
\end{equation*}
$$

Proof. Averaging the previous result over all $S^{2} H^{2}$ possibilities for the quadruple $\left(s, s^{\prime}, h, h^{\prime}\right)$, we get $2 m-1 \leq \frac{a}{S^{2}}+\frac{b}{S H}+\frac{d}{H^{2}}$. Due to the symmetry between rows and columns, the same estimate holds with $b$ replaced by $c$. Averaging these two inequalities we arrive at our claim.

Now let us multiply (2) by $H^{2}$, multiply (11) by $\left(2 S H-H^{2}\right)$ and add them; we get
$H^{2}(2 m-1)+\left(2 S H-H^{2}\right) m \leq a \cdot \frac{H^{2}+2 S H-H^{2}}{S^{2}}+(b+c) \frac{H^{2}+2 S H-H^{2}}{2 S H}+d=a \cdot \frac{2 H}{S}+b+c+d$.
The left-hand side is exactly $H^{2}(m-1)+2 S H m$, while the right-hand side does not exceed $a+b+c+d$ since $2 H \leq S$. Hence we come to the desired inequality.

Comment 1. Claim 2 is the key difference between the two solutions, because it allows to get rid of the notions of rich and poor cells. However, one may prove it by the "strawberry method" as well. It suffices to put a strawberry on each cell which is bad for an s-row, and a strawberry on each cell which is bad for an $h^{\prime}$-column. Then each cell would contain not more than one strawberry.

Comment 2. Both solutions above work if the residue of the table size $T$ modulo the napkin size $R$ is at least $\frac{2}{3} R$, or equivalently if $T=S m+H(m-1)$ and $R=S+H$ for some positive integers $S, H$, $m$ such that $S \geq 2 H$. Here we discuss all other possible combinations.

Case 1. If $2 H \geq S \geq H / 2$, then the sharp bound for the number of vicious cells is $m S^{2}+(m-1) H^{2}$; it can be obtained by the same methods as in any of the solutions. To obtain an example showing that the bound is sharp, one may simply remove the napkins of the third kind from the example in Solution 1 (with an obvious change in the numbers).

Case 2. If $2 S \leq H$, the situation is more difficult. If $(S+H)^{2}>2 H^{2}$, then the answer and the example are the same as in the previous case; otherwise the answer is $(2 m-1) S^{2}+2 S H(m-1)$, and the example is provided simply by $(m-1)^{2}$ nonintersecting napkins.

Now we sketch the proof of both estimates for Case 2. We introduce a more appropriate notation based on that from Solution 2. Denote by $a_{-}$and $a_{+}$the number of cells of class $A$ that contain the number which is strictly less than $M$ and strictly greater than $M$, respectively. The numbers $b_{ \pm}, c_{ \pm}$, and $d_{ \pm}$are defined in a similar way. One may notice that the proofs of Claim 1 and Claims 2, 3 lead in fact to the inequalities

$$
m-1 \leq \frac{b_{-}+c_{-}}{2 S H}+\frac{d_{+}}{H^{2}} \quad \text { and } \quad 2 m-1 \leq \frac{a}{S^{2}}+\frac{b_{+}+c_{+}}{2 S H}+\frac{d_{+}}{H^{2}}
$$

(to obtain the first one, one needs to look at the big lines instead of the small ones). Combining these inequalities, one may obtain the desired estimates.

These estimates can also be proved in some different ways, e.g. without distinguishing rich and poor cells.

## G1

Let $A B C$ be an acute triangle. Let $\omega$ be a circle whose center $L$ lies on the side $B C$. Suppose that $\omega$ is tangent to $A B$ at $B^{\prime}$ and to $A C$ at $C^{\prime}$. Suppose also that the circumcenter $O$ of the triangle $A B C$ lies on the shorter arc $B^{\prime} C^{\prime}$ of $\omega$. Prove that the circumcircle of $A B C$ and $\omega$ meet at two points.

Solution. The point $B^{\prime}$, being the perpendicular foot of $L$, is an interior point of side $A B$. Analogously, $C^{\prime}$ lies in the interior of $A C$. The point $O$ is located inside the triangle $A B^{\prime} C^{\prime}$, hence $\angle C O B<\angle C^{\prime} O B^{\prime}$.


Let $\alpha=\angle C A B$. The angles $\angle C A B$ and $\angle C^{\prime} O B^{\prime}$ are inscribed into the two circles with centers $O$ and $L$, respectively, so $\angle C O B=2 \angle C A B=2 \alpha$ and $2 \angle C^{\prime} O B^{\prime}=360^{\circ}-\angle C^{\prime} L B^{\prime}$. From the kite $A B^{\prime} L C^{\prime}$ we have $\angle C^{\prime} L B^{\prime}=180^{\circ}-\angle C^{\prime} A B^{\prime}=180^{\circ}-\alpha$. Combining these, we get

$$
2 \alpha=\angle C O B<\angle C^{\prime} O B^{\prime}=\frac{360^{\circ}-\angle C^{\prime} L B^{\prime}}{2}=\frac{360^{\circ}-\left(180^{\circ}-\alpha\right)}{2}=90^{\circ}+\frac{\alpha}{2},
$$

so

$$
\alpha<60^{\circ} .
$$

Let $O^{\prime}$ be the reflection of $O$ in the line $B C$. In the quadrilateral $A B O^{\prime} C$ we have

$$
\angle C O^{\prime} B+\angle C A B=\angle C O B+\angle C A B=2 \alpha+\alpha<180^{\circ},
$$

so the point $O^{\prime}$ is outside the circle $A B C$. Hence, $O$ and $O^{\prime}$ are two points of $\omega$ such that one of them lies inside the circumcircle, while the other one is located outside. Therefore, the two circles intersect.

Comment. There are different ways of reducing the statement of the problem to the case $\alpha<60^{\circ}$. E.g., since the point $O$ lies in the interior of the isosceles triangle $A B^{\prime} C^{\prime}$, we have $O A<A B^{\prime}$. So, if $A B^{\prime} \leq 2 L B^{\prime}$ then $O A<2 L O$, which means that $\omega$ intersects the circumcircle of $A B C$. Hence the only interesting case is $A B^{\prime}>2 L B^{\prime}$, and this condition implies $\angle C A B=2 \angle B^{\prime} A L<2 \cdot 30^{\circ}=60^{\circ}$.

## G2

Let $A_{1} A_{2} A_{3} A_{4}$ be a non-cyclic quadrilateral. Let $O_{1}$ and $r_{1}$ be the circumcenter and the circumradius of the triangle $A_{2} A_{3} A_{4}$. Define $O_{2}, O_{3}, O_{4}$ and $r_{2}, r_{3}, r_{4}$ in a similar way. Prove that

$$
\frac{1}{O_{1} A_{1}^{2}-r_{1}^{2}}+\frac{1}{O_{2} A_{2}^{2}-r_{2}^{2}}+\frac{1}{O_{3} A_{3}^{2}-r_{3}^{2}}+\frac{1}{O_{4} A_{4}^{2}-r_{4}^{2}}=0
$$

Solution 1. Let $M$ be the point of intersection of the diagonals $A_{1} A_{3}$ and $A_{2} A_{4}$. On each diagonal choose a direction and let $x, y, z$, and $w$ be the signed distances from $M$ to the points $A_{1}, A_{2}, A_{3}$, and $A_{4}$, respectively.

Let $\omega_{1}$ be the circumcircle of the triangle $A_{2} A_{3} A_{4}$ and let $B_{1}$ be the second intersection point of $\omega_{1}$ and $A_{1} A_{3}$ (thus, $B_{1}=A_{3}$ if and only if $A_{1} A_{3}$ is tangent to $\omega_{1}$ ). Since the expression $O_{1} A_{1}^{2}-r_{1}^{2}$ is the power of the point $A_{1}$ with respect to $\omega_{1}$, we get

$$
O_{1} A_{1}^{2}-r_{1}^{2}=A_{1} B_{1} \cdot A_{1} A_{3} .
$$

On the other hand, from the equality $M B_{1} \cdot M A_{3}=M A_{2} \cdot M A_{4}$ we obtain $M B_{1}=y w / z$. Hence, we have

$$
O_{1} A_{1}^{2}-r_{1}^{2}=\left(\frac{y w}{z}-x\right)(z-x)=\frac{z-x}{z}(y w-x z) .
$$

Substituting the analogous expressions into the sought sum we get

$$
\sum_{i=1}^{4} \frac{1}{O_{i} A_{i}^{2}-r_{i}^{2}}=\frac{1}{y w-x z}\left(\frac{z}{z-x}-\frac{w}{w-y}+\frac{x}{x-z}-\frac{y}{y-w}\right)=0
$$

as desired.

Comment. One might reformulate the problem by assuming that the quadrilateral $A_{1} A_{2} A_{3} A_{4}$ is convex. This should not really change the difficulty, but proofs that distinguish several cases may become shorter.

Solution 2. Introduce a Cartesian coordinate system in the plane. Every circle has an equation of the form $p(x, y)=x^{2}+y^{2}+l(x, y)=0$, where $l(x, y)$ is a polynomial of degree at most 1 . For any point $A=\left(x_{A}, y_{A}\right)$ we have $p\left(x_{A}, y_{A}\right)=d^{2}-r^{2}$, where $d$ is the distance from $A$ to the center of the circle and $r$ is the radius of the circle.

For each $i$ in $\{1,2,3,4\}$ let $p_{i}(x, y)=x^{2}+y^{2}+l_{i}(x, y)=0$ be the equation of the circle with center $O_{i}$ and radius $r_{i}$ and let $d_{i}$ be the distance from $A_{i}$ to $O_{i}$. Consider the equation

$$
\begin{equation*}
\sum_{i=1}^{4} \frac{p_{i}(x, y)}{d_{i}^{2}-r_{i}^{2}}=1 \tag{1}
\end{equation*}
$$

Since the coordinates of the points $A_{1}, A_{2}, A_{3}$, and $A_{4}$ satisfy (1) but these four points do not lie on a circle or on an line, equation (11) defines neither a circle, nor a line. Hence, the equation is an identity and the coefficient of the quadratic term $x^{2}+y^{2}$ also has to be zero, i.e.

$$
\sum_{i=1}^{4} \frac{1}{d_{i}^{2}-r_{i}^{2}}=0
$$

Comment. Using the determinant form of the equation of the circle through three given points, the same solution can be formulated as follows.

For $i=1,2,3,4$ let $\left(u_{i}, v_{i}\right)$ be the coordinates of $A_{i}$ and define

$$
\Delta=\left|\begin{array}{llll}
u_{1}^{2}+v_{1}^{2} & u_{1} & v_{1} & 1 \\
u_{2}^{2}+v_{2}^{2} & u_{2} & v_{2} & 1 \\
u_{3}^{2}+v_{3}^{2} & u_{3} & v_{3} & 1 \\
u_{4}^{2}+v_{4}^{2} & u_{4} & v_{4} & 1
\end{array}\right| \quad \text { and } \quad \Delta_{i}=\left|\begin{array}{ccc}
u_{i+1} & v_{i+1} & 1 \\
u_{i+2} & v_{i+2} & 1 \\
u_{i+3} & v_{i+3} & 1
\end{array}\right|,
$$

where $i+1, i+2$, and $i+3$ have to be read modulo 4 as integers in the set $\{1,2,3,4\}$.
Expanding $\left|\begin{array}{llll}u_{1} & v_{1} & 1 & 1 \\ u_{2} & v_{2} & 1 & 1 \\ u_{3} & v_{3} & 1 & 1 \\ u_{4} & v_{4} & 1 & 1\end{array}\right|=0$ along the third column, we get $\Delta_{1}-\Delta_{2}+\Delta_{3}-\Delta_{4}=0$.
The circle through $A_{i+1}, A_{i+2}$, and $A_{i+3}$ is given by the equation

$$
\frac{1}{\Delta_{i}}\left|\begin{array}{cccc}
x^{2}+y^{2} & x & y & 1  \tag{2}\\
u_{i+1}^{2}+v_{i+1}^{2} & u_{i+1} & v_{i+1} & 1 \\
u_{i+2}^{2}+v_{i+2}^{2} & u_{i+2} & v_{i+2} & 1 \\
u_{i+3}^{2}+v_{i+3}^{2} & u_{i+3} & v_{i+3} & 1
\end{array}\right|=0
$$

On the left-hand side, the coefficient of $x^{2}+y^{2}$ is equal to 1 . Substituting $\left(u_{i}, v_{i}\right)$ for $(x, y)$ in (2) we obtain the power of point $A_{i}$ with respect to the circle through $A_{i+1}, A_{i+2}$, and $A_{i+3}$ :

$$
d_{i}^{2}-r_{i}^{2}=\frac{1}{\Delta_{i}}\left|\begin{array}{cccc}
u_{i}^{2}+v_{i}^{2} & u_{i} & v_{i} & 1 \\
u_{i+1}^{2}+v_{i+1}^{2} & u_{i+1} & v_{i+1} & 1 \\
u_{i+2}^{2}+v_{i+2}^{2} & u_{i+2} & v_{i+2} & 1 \\
u_{i+3}^{2}+v_{i+3}^{2} & u_{i+3} & v_{i+3} & 1
\end{array}\right|=(-1)^{i+1} \frac{\Delta}{\Delta_{i}} .
$$

Thus, we have

$$
\sum_{i=1}^{4} \frac{1}{d_{i}^{2}-r_{i}^{2}}=\frac{\Delta_{1}-\Delta_{2}+\Delta_{3}-\Delta_{4}}{\Delta}=0 .
$$

## G3

Let $A B C D$ be a convex quadrilateral whose sides $A D$ and $B C$ are not parallel. Suppose that the circles with diameters $A B$ and $C D$ meet at points $E$ and $F$ inside the quadrilateral. Let $\omega_{E}$ be the circle through the feet of the perpendiculars from $E$ to the lines $A B, B C$, and $C D$. Let $\omega_{F}$ be the circle through the feet of the perpendiculars from $F$ to the lines $C D, D A$, and $A B$. Prove that the midpoint of the segment $E F$ lies on the line through the two intersection points of $\omega_{E}$ and $\omega_{F}$.

Solution. Denote by $P, Q, R$, and $S$ the projections of $E$ on the lines $D A, A B, B C$, and $C D$ respectively. The points $P$ and $Q$ lie on the circle with diameter $A E$, so $\angle Q P E=\angle Q A E$; analogously, $\angle Q R E=\angle Q B E$. So $\angle Q P E+\angle Q R E=\angle Q A E+\angle Q B E=90^{\circ}$. By similar reasons, we have $\angle S P E+\angle S R E=90^{\circ}$, hence we get $\angle Q P S+\angle Q R S=90^{\circ}+90^{\circ}=180^{\circ}$, and the quadrilateral $P Q R S$ is inscribed in $\omega_{E}$. Analogously, all four projections of $F$ onto the sides of $A B C D$ lie on $\omega_{F}$.

Denote by $K$ the meeting point of the lines $A D$ and $B C$. Due to the arguments above, there is no loss of generality in assuming that $A$ lies on segment $D K$. Suppose that $\angle C K D \geq 90^{\circ}$; then the circle with diameter $C D$ covers the whole quadrilateral $A B C D$, so the points $E, F$ cannot lie inside this quadrilateral. Hence our assumption is wrong. Therefore, the lines $E P$ and $B C$ intersect at some point $P^{\prime}$, while the lines $E R$ and $A D$ intersect at some point $R^{\prime}$.


Figure 1
We claim that the points $P^{\prime}$ and $R^{\prime}$ also belong to $\omega_{E}$. Since the points $R, E, Q, B$ are concyclic, $\angle Q R K=\angle Q E B=90^{\circ}-\angle Q B E=\angle Q A E=\angle Q P E$. So $\angle Q R K=\angle Q P P^{\prime}$, which means that the point $P^{\prime}$ lies on $\omega_{E}$. Analogously, $R^{\prime}$ also lies on $\omega_{E}$.

In the same manner, denote by $M$ and $N$ the projections of $F$ on the lines $A D$ and $B C$
respectively, and let $M^{\prime}=F M \cap B C, N^{\prime}=F N \cap A D$. By the same arguments, we obtain that the points $M^{\prime}$ and $N^{\prime}$ belong to $\omega_{F}$.


Figure 2
Now we concentrate on Figure 2, where all unnecessary details are removed. Let $U=N N^{\prime} \cap$ $P P^{\prime}, V=M M^{\prime} \cap R R^{\prime}$. Due to the right angles at $N$ and $P$, the points $N, N^{\prime}, P, P^{\prime}$ are concyclic, so $U N \cdot U N^{\prime}=U P \cdot U P^{\prime}$ which means that $U$ belongs to the radical axis $g$ of the circles $\omega_{E}$ and $\omega_{F}$. Analogously, $V$ also belongs to $g$.
Finally, since $E U F V$ is a parallelogram, the radical axis $U V$ of $\omega_{E}$ and $\omega_{F}$ bisects $E F$.

## G4

Let $A B C$ be an acute triangle with circumcircle $\Omega$. Let $B_{0}$ be the midpoint of $A C$ and let $C_{0}$ be the midpoint of $A B$. Let $D$ be the foot of the altitude from $A$, and let $G$ be the centroid of the triangle $A B C$. Let $\omega$ be a circle through $B_{0}$ and $C_{0}$ that is tangent to the circle $\Omega$ at a point $X \neq A$. Prove that the points $D, G$, and $X$ are collinear.

Solution 1. If $A B=A C$, then the statement is trivial. So without loss of generality we may assume $A B<A C$. Denote the tangents to $\Omega$ at points $A$ and $X$ by $a$ and $x$, respectively.

Let $\Omega_{1}$ be the circumcircle of triangle $A B_{0} C_{0}$. The circles $\Omega$ and $\Omega_{1}$ are homothetic with center $A$, so they are tangent at $A$, and $a$ is their radical axis. Now, the lines $a, x$, and $B_{0} C_{0}$ are the three radical axes of the circles $\Omega, \Omega_{1}$, and $\omega$. Since $a \nmid B_{0} C_{0}$, these three lines are concurrent at some point $W$.

The points $A$ and $D$ are symmetric with respect to the line $B_{0} C_{0}$; hence $W X=W A=W D$. This means that $W$ is the center of the circumcircle $\gamma$ of triangle $A D X$. Moreover, we have $\angle W A O=\angle W X O=90^{\circ}$, where $O$ denotes the center of $\Omega$. Hence $\angle A W X+\angle A O X=180^{\circ}$.


Denote by $T$ the second intersection point of $\Omega$ and the line $D X$. Note that $O$ belongs to $\Omega_{1}$. Using the circles $\gamma$ and $\Omega$, we find $\angle D A T=\angle A D X-\angle A T D=\frac{1}{2}\left(360^{\circ}-\angle A W X\right)-\frac{1}{2} \angle A O X=$ $180^{\circ}-\frac{1}{2}(\angle A W X+\angle A O X)=90^{\circ}$. So, $A D \perp A T$, and hence $A T \| B C$. Thus, $A T C B$ is an isosceles trapezoid inscribed in $\Omega$.

Denote by $A_{0}$ the midpoint of $B C$, and consider the image of $A T C B$ under the homothety $h$ with center $G$ and factor $-\frac{1}{2}$. We have $h(A)=A_{0}, h(B)=B_{0}$, and $h(C)=C_{0}$. From the
symmetry about $B_{0} C_{0}$, we have $\angle T C B=\angle C B A=\angle B_{0} C_{0} A=\angle D C_{0} B_{0}$. Using $A T \| D A_{0}$, we conclude $h(T)=D$. Hence the points $D, G$, and $T$ are collinear, and $X$ lies on the same line.

Solution 2. We define the points $A_{0}, O$, and $W$ as in the previous solution and we concentrate on the case $A B<A C$. Let $Q$ be the perpendicular projection of $A_{0}$ on $B_{0} C_{0}$.

Since $\angle W A O=\angle W Q O=\angle O X W=90^{\circ}$, the five points $A, W, X, O$, and $Q$ lie on a common circle. Furthermore, the reflections with respect to $B_{0} C_{0}$ and $O W$ map $A$ to $D$ and $X$, respectively. For these reasons, we have

$$
\angle W Q D=\angle A Q W=\angle A X W=\angle W A X=\angle W Q X
$$

Thus the three points $Q, D$, and $X$ lie on a common line, say $\ell$.


To complete the argument, we note that the homothety centered at $G$ sending the triangle $A B C$ to the triangle $A_{0} B_{0} C_{0}$ maps the altitude $A D$ to the altitude $A_{0} Q$. Therefore it maps $D$ to $Q$, so the points $D, G$, and $Q$ are collinear. Hence $G$ lies on $\ell$ as well.

Comment. There are various other ways to prove the collinearity of $Q, D$, and $X$ obtained in the middle part of Solution 2. Introduce for instance the point $J$ where the lines $A W$ and $B C$ intersect. Then the four points $A, D, X$, and $J$ lie at the same distance from $W$, so the quadrilateral $A D X J$ is cyclic. In combination with the fact that $A W X Q$ is cyclic as well, this implies

$$
\angle J D X=\angle J A X=\angle W A X=\angle W Q X
$$

Since $B C \| W Q$, it follows that $Q, D$, and $X$ are indeed collinear.

## G5

Let $A B C$ be a triangle with incenter $I$ and circumcircle $\omega$. Let $D$ and $E$ be the second intersection points of $\omega$ with the lines $A I$ and $B I$, respectively. The chord $D E$ meets $A C$ at a point $F$, and $B C$ at a point $G$. Let $P$ be the intersection point of the line through $F$ parallel to $A D$ and the line through $G$ parallel to $B E$. Suppose that the tangents to $\omega$ at $A$ and at $B$ meet at a point $K$. Prove that the three lines $A E, B D$, and $K P$ are either parallel or concurrent.

Solution 1. Since

$$
\angle I A F=\angle D A C=\angle B A D=\angle B E D=\angle I E F
$$

the quadrilateral $A I F E$ is cyclic. Denote its circumcircle by $\omega_{1}$. Similarly, the quadrilateral $B D G I$ is cyclic; denote its circumcircle by $\omega_{2}$.

The line $A E$ is the radical axis of $\omega$ and $\omega_{1}$, and the line $B D$ is the radical axis of $\omega$ and $\omega_{2}$. Let $t$ be the radical axis of $\omega_{1}$ and $\omega_{2}$. These three lines meet at the radical center of the three circles, or they are parallel to each other. We will show that $t$ is in fact the line PK.

Let $L$ be the second intersection point of $\omega_{1}$ and $\omega_{2}$, so $t=I L$. (If the two circles are tangent to each other then $L=I$ and $t$ is the common tangent.)


Let the line $t$ meet the circumcircles of the triangles $A B L$ and $F G L$ at $K^{\prime} \neq L$ and $P^{\prime} \neq L$, respectively. Using oriented angles we have

$$
\angle\left(A B, B K^{\prime}\right)=\angle\left(A L, L K^{\prime}\right)=\angle(A L, L I)=\angle(A E, E I)=\angle(A E, E B)=\angle(A B, B K),
$$

so $B K^{\prime} \| B K$. Similarly we have $A K^{\prime} \| A K$, and therefore $K^{\prime}=K$. Next, we have

$$
\angle\left(P^{\prime} F, F G\right)=\angle\left(P^{\prime} L, L G\right)=\angle(I L, L G)=\angle(I D, D G)=\angle(A D, D E)=\angle(P F, F G),
$$

hence $P^{\prime} F \| P F$ and similarly $P^{\prime} G \| P G$. Therefore $P^{\prime}=P$. This means that $t$ passes through $K$ and $P$, which finishes the proof.

Solution 2. Let $M$ be the intersection point of the tangents to $\omega$ at $D$ and $E$, and let the lines $A E$ and $B D$ meet at $T$; if $A E$ and $B D$ are parallel, then let $T$ be their common ideal point. It is well-known that the points $K$ and $M$ lie on the line $T I$ (as a consequence of Pascal's theorem, applied to the inscribed degenerate hexagons $A A D B B E$ and $A D D B E E$ ).

The lines $A D$ and $B E$ are the angle bisectors of the angles $\angle C A B$ and $\angle A B C$, respectively, so $D$ and $E$ are the midpoints of the arcs $B C$ and $C A$ of the circle $\omega$, respectively. Hence, $D M$ is parallel to $B C$ and $E M$ is parallel to $A C$.

Apply Pascal's theorem to the degenerate hexagon $C A D D E B$. By the theorem, the points $C A \cap D E=F, A D \cap E B=I$ and the common ideal point of lines $D M$ and $B C$ are collinear, therefore $F I$ is parallel to $B C$ and $D M$. Analogously, the line $G I$ is parallel to $A C$ and $E M$.


Now consider the homothety with scale factor $-\frac{F G}{E D}$ which takes $E$ to $G$ and $D$ to $F$. Since the triangles $E D M$ and $G F I$ have parallel sides, the homothety takes $M$ to $I$. Similarly, since the triangles $D E I$ and $F G P$ have parallel sides, the homothety takes $I$ to $P$. Hence, the points $M, I, P$ and the homothety center $H$ must lie on the same line. Therefore, the point $P$ also lies on the line TKIM.

Comment. One may prove that $I F \| B C$ and $I G \| A C$ in a more elementary way. Since $\angle A D E=$ $\angle E D C$ and $\angle D E B=\angle C E D$, the points $I$ and $C$ are symmetric about $D E$. Moreover, since the $\operatorname{arcs} A E$ and $E C$ are equal and the arcs $C D$ and $D B$ are equal, we have $\angle C F G=\angle F G C$, so the triangle $C F G$ is isosceles. Hence, the quadrilateral $I F C G$ is a rhombus.

## G6

Let $A B C$ be a triangle with $A B=A C$, and let $D$ be the midpoint of $A C$. The angle bisector of $\angle B A C$ intersects the circle through $D, B$, and $C$ in a point $E$ inside the triangle $A B C$. The line $B D$ intersects the circle through $A, E$, and $B$ in two points $B$ and $F$. The lines $A F$ and $B E$ meet at a point $I$, and the lines $C I$ and $B D$ meet at a point $K$. Show that $I$ is the incenter of triangle $K A B$.

Solution 1. Let $D^{\prime}$ be the midpoint of the segment $A B$, and let $M$ be the midpoint of $B C$. By symmetry at line $A M$, the point $D^{\prime}$ has to lie on the circle $B C D$. Since the $\operatorname{arcs} D^{\prime} E$ and $E D$ of that circle are equal, we have $\angle A B I=\angle D^{\prime} B E=\angle E B D=I B K$, so $I$ lies on the angle bisector of $\angle A B K$. For this reason it suffices to prove in the sequel that the ray $A I$ bisects the angle $\angle B A K$.

From

$$
\angle D F A=180^{\circ}-\angle B F A=180^{\circ}-\angle B E A=\angle M E B=\frac{1}{2} \angle C E B=\frac{1}{2} \angle C D B
$$

we derive $\angle D F A=\angle D A F$ so the triangle $A F D$ is isosceles with $A D=D F$.


Applying Menelaus's theorem to the triangle $A D F$ with respect to the line $C I K$, and applying the angle bisector theorem to the triangle $A B F$, we infer

$$
1=\frac{A C}{C D} \cdot \frac{D K}{K F} \cdot \frac{F I}{I A}=2 \cdot \frac{D K}{K F} \cdot \frac{B F}{A B}=2 \cdot \frac{D K}{K F} \cdot \frac{B F}{2 \cdot A D}=\frac{D K}{K F} \cdot \frac{B F}{A D}
$$

and therefore

$$
\frac{B D}{A D}=\frac{B F+F D}{A D}=\frac{B F}{A D}+1=\frac{K F}{D K}+1=\frac{D F}{D K}=\frac{A D}{D K}
$$

It follows that the triangles $A D K$ and $B D A$ are similar, hence $\angle D A K=\angle A B D$. Then

$$
\angle I A B=\angle A F D-\angle A B D=\angle D A F-\angle D A K=\angle K A I
$$

shows that the point $K$ is indeed lying on the angle bisector of $\angle B A K$.

Solution 2. It can be shown in the same way as in the first solution that $I$ lies on the angle bisector of $\angle A B K$. Here we restrict ourselves to proving that $K I$ bisects $\angle A K B$.


Denote the circumcircle of triangle $B C D$ and its center by $\omega_{1}$ and by $O_{1}$, respectively. Since the quadrilateral $A B F E$ is cyclic, we have $\angle D F E=\angle B A E=\angle D A E$. By the same reason, we have $\angle E A F=\angle E B F=\angle A B E=\angle A F E$. Therefore $\angle D A F=\angle D F A$, and hence $D F=D A=D C$. So triangle $A F C$ is inscribed in a circle $\omega_{2}$ with center $D$.

Denote the circumcircle of triangle $A B D$ by $\omega_{3}$, and let its center be $O_{3}$. Since the $\operatorname{arcs} B E$ and $E C$ of circle $\omega_{1}$ are equal, and the triangles $A D E$ and $F D E$ are congruent, we have $\angle A O_{1} B=2 \angle B D E=\angle B D A$, so $O_{1}$ lies on $\omega_{3}$. Hence $\angle O_{3} O_{1} D=\angle O_{3} D O_{1}$.

The line $B D$ is the radical axis of $\omega_{1}$ and $\omega_{3}$. Point $C$ belongs to the radical axis of $\omega_{1}$ and $\omega_{2}$, and $I$ also belongs to it since $A I \cdot I F=B I \cdot I E$. Hence $K=B D \cap C I$ is the radical center of $\omega_{1}$, $\omega_{2}$, and $\omega_{3}$, and $A K$ is the radical axis of $\omega_{2}$ and $\omega_{3}$. Now, the radical axes $A K, B K$ and $I K$ are perpendicular to the central lines $O_{3} D, O_{3} O_{1}$ and $O_{1} D$, respectively. By $\angle O_{3} O_{1} D=\angle O_{3} D O_{1}$, we get that $K I$ is the angle bisector of $\angle A K B$.

Solution 3. Again, let $M$ be the midpoint of $B C$. As in the previous solutions, we can deduce $\angle A B I=\angle I B K$. We show that the point $I$ lies on the angle bisector of $\angle K A B$.

Let $G$ be the intersection point of the circles $A F C$ and $B C D$, different from $C$. The lines
$C G, A F$, and $B E$ are the radical axes of the three circles $A G F C, C D B$, and $A B F E$, so $I=A F \cap B E$ is the radical center of the three circles and $C G$ also passes through $I$.


The angle between line $D E$ and the tangent to the circle $B C D$ at $E$ is equal to $\angle E B D=$ $\angle E A F=\angle A B E=\angle A F E$. As the tangent at $E$ is perpendicular to $A M$, the line $D E$ is perpendicular to $A F$. The triangle $A F E$ is isosceles, so $D E$ is the perpendicular bisector of $A F$ and thus $A D=D F$. Hence, the point $D$ is the center of the circle $A F C$, and this circle passes through $M$ as well since $\angle A M C=90^{\circ}$.

Let $B^{\prime}$ be the reflection of $B$ in the point $D$, so $A B C B^{\prime}$ is a parallelogram. Since $D C=D G$ we have $\angle G C D=\angle D B C=\angle K B^{\prime} A$. Hence, the quadrilateral $A K C B^{\prime}$ is cyclic and thus $\angle C A K=\angle C B^{\prime} K=\angle A B D=2 \angle M A I$. Then

$$
\angle I A B=\angle M A B-\angle M A I=\frac{1}{2} \angle C A B-\frac{1}{2} \angle C A K=\frac{1}{2} \angle K A B
$$

and therefore $A I$ is the angle bisector of $\angle K A B$.

## G7

Let $A B C D E F$ be a convex hexagon all of whose sides are tangent to a circle $\omega$ with center $O$. Suppose that the circumcircle of triangle $A C E$ is concentric with $\omega$. Let $J$ be the foot of the perpendicular from $B$ to $C D$. Suppose that the perpendicular from $B$ to $D F$ intersects the line $E O$ at a point $K$. Let $L$ be the foot of the perpendicular from $K$ to $D E$. Prove that $D J=D L$.

Solution 1. Since $\omega$ and the circumcircle of triangle $A C E$ are concentric, the tangents from $A$, $C$, and $E$ to $\omega$ have equal lengths; that means that $A B=B C, C D=D E$, and $E F=F A$. Moreover, we have $\angle B C D=\angle D E F=\angle F A B$.


Consider the rotation around point $D$ mapping $C$ to $E$; let $B^{\prime}$ and $L^{\prime}$ be the images of the points $B$ and $J$, respectively, under this rotation. Then one has $D J=D L^{\prime}$ and $B^{\prime} L^{\prime} \perp D E$; moreover, the triangles $B^{\prime} E D$ and $B C D$ are congruent. Since $\angle D E O<90^{\circ}$, the lines $E O$ and $B^{\prime} L^{\prime}$ intersect at some point $K^{\prime}$. We intend to prove that $K^{\prime} B \perp D F$; this would imply $K=K^{\prime}$, therefore $L=L^{\prime}$, which proves the problem statement.

Analogously, consider the rotation around $F$ mapping $A$ to $E$; let $B^{\prime \prime}$ be the image of $B$ under this rotation. Then the triangles $F A B$ and $F E B^{\prime \prime}$ are congruent. We have $E B^{\prime \prime}=A B=B C=$ $E B^{\prime}$ and $\angle F E B^{\prime \prime}=\angle F A B=\angle B C D=\angle D E B^{\prime}$, so the points $B^{\prime}$ and $B^{\prime \prime}$ are symmetrical with respect to the angle bisector $E O$ of $\angle D E F$. So, from $K^{\prime} B^{\prime} \perp D E$ we get $K^{\prime} B^{\prime \prime} \perp E F$.

From these two relations we obtain

$$
K^{\prime} D^{2}-K^{\prime} E^{2}=B^{\prime} D^{2}-B^{\prime} E^{2} \quad \text { and } \quad K^{\prime} E^{2}-K^{\prime} F^{2}=B^{\prime \prime} E^{2}-B^{\prime \prime} F^{2} .
$$

Adding these equalities and taking into account that $B^{\prime} E=B^{\prime \prime} E$ we obtain

$$
K^{\prime} D^{2}-K^{\prime} F^{2}=B^{\prime} D^{2}-B^{\prime \prime} F^{2}=B D^{2}-B F^{2}
$$

which means exactly that $K^{\prime} B \perp D F$.

Comment. There are several variations of this solution. For instance, let us consider the intersection point $M$ of the lines $B J$ and $O C$. Define the point $K^{\prime}$ as the reflection of $M$ in the line $D O$. Then one has

$$
D K^{\prime 2}-D B^{2}=D M^{2}-D B^{2}=C M^{2}-C B^{2} .
$$

Next, consider the rotation around $O$ which maps $C M$ to $E K^{\prime}$. Let $P$ be the image of $B$ under this rotation; so $P$ lies on $E D$. Then $E F \perp K^{\prime} P$, so

$$
C M^{2}-C B^{2}=E K^{\prime 2}-E P^{2}=F K^{\prime 2}-F P^{2}=F K^{\prime 2}-F B^{2},
$$

since the triangles $F E P$ and $F A B$ are congruent.

Solution 2. Let us denote the points of tangency of $A B, B C, C D, D E, E F$, and $F A$ to $\omega$ by $R, S, T, U, V$, and $W$, respectively. As in the previous solution, we mention that $A R=$ $A W=C S=C T=E U=E V$.

The reflection in the line $B O$ maps $R$ to $S$, therefore $A$ to $C$ and thus also $W$ to $T$. Hence, both lines $R S$ and $W T$ are perpendicular to $O B$, therefore they are parallel. On the other hand, the lines $U V$ and $W T$ are not parallel, since otherwise the hexagon $A B C D E F$ is symmetric with respect to the line $B O$ and the lines defining the point $K$ coincide, which contradicts the conditions of the problem. Therefore we can consider the intersection point $Z$ of $U V$ and $W T$.


Next, we recall a well-known fact that the points $D, F, Z$ are collinear. Actually, $D$ is the pole of the line $U T, F$ is the pole of $V W$, and $Z=T W \cap U V$; so all these points belong to the polar line of $T U \cap V W$.

Now, we put $O$ into the origin, and identify each point (say $X$ ) with the vector $\overrightarrow{O X}$. So, from now on all the products of points refer to the scalar products of the corresponding vectors.

Since $O K \perp U Z$ and $O B \perp T Z$, we have $K \cdot(Z-U)=0=B \cdot(Z-T)$. Next, the condition $B K \perp D Z$ can be written as $K \cdot(D-Z)=B \cdot(D-Z)$. Adding these two equalities we get

$$
K \cdot(D-U)=B \cdot(D-T) .
$$

By symmetry, we have $D \cdot(D-U)=D \cdot(D-T)$. Subtracting this from the previous equation, we obtain $(K-D) \cdot(D-U)=(B-D) \cdot(D-T)$ and rewrite it in vector form as

$$
\overrightarrow{D K} \cdot \overrightarrow{U D}=\overrightarrow{D B} \cdot \overrightarrow{T D}
$$

Finally, projecting the vectors $\overrightarrow{D K}$ and $\overrightarrow{D B}$ onto the lines $U D$ and $T D$ respectively, we can rewrite this equality in terms of segment lengths as $D L \cdot U D=D J \cdot T D$, thus $D L=D J$.

Comment. The collinearity of $Z, F$, and $D$ may be shown in various more elementary ways. For instance, applying the sine theorem to the triangles $D T Z$ and $D U Z$, one gets $\frac{\sin \angle D Z T}{\sin \angle D Z U}=\frac{\sin \angle D T Z}{\sin \angle D U Z}$; analogously, $\frac{\sin \angle F Z W}{\sin \angle F Z V}=\frac{\sin \angle F W Z}{\sin \angle F V Z}$. The right-hand sides are equal, hence so are the left-hand sides, which implies the collinearity of the points $D, F$, and $Z$.

There also exist purely synthetic proofs of this fact. E.g., let $Q$ be the point of intersection of the circumcircles of the triangles $Z T V$ and $Z W U$ different from $Z$. Then $Q Z$ is the bisector of $\angle V Q W$ since $\angle V Q Z=\angle V T Z=\angle V U W=\angle Z Q W$. Moreover, all these angles are equal to $\frac{1}{2} \angle V O W$, so $\angle V Q W=\angle V O W$, hence the quadrilateral $V W O Q$ is cyclic. On the other hand, the points $O$, $V, W$ lie on the circle with diameter $O F$ due to the right angles; so $Q$ also belongs to this circle. Since $F V=F W, Q F$ is also the bisector of $\angle V Q W$, so $F$ lies on $Q Z$. Analogously, $D$ lies on the same line.

## G8

Let $A B C$ be an acute triangle with circumcircle $\omega$. Let $t$ be a tangent line to $\omega$. Let $t_{a}, t_{b}$, and $t_{c}$ be the lines obtained by reflecting $t$ in the lines $B C, C A$, and $A B$, respectively. Show that the circumcircle of the triangle determined by the lines $t_{a}, t_{b}$, and $t_{c}$ is tangent to the circle $\omega$.

To avoid a large case distinction, we will use the notion of oriented angles. Namely, for two lines $\ell$ and $m$, we denote by $\angle(\ell, m)$ the angle by which one may rotate $\ell$ anticlockwise to obtain a line parallel to $m$. Thus, all oriented angles are considered modulo $180^{\circ}$.


Solution 1. Denote by $T$ the point of tangency of $t$ and $\omega$. Let $A^{\prime}=t_{b} \cap t_{c}, B^{\prime}=t_{a} \cap t_{c}$, $C^{\prime}=t_{a} \cap t_{b}$. Introduce the point $A^{\prime \prime}$ on $\omega$ such that $T A=A A^{\prime \prime}\left(A^{\prime \prime} \neq T\right.$ unless $T A$ is a diameter). Define the points $B^{\prime \prime}$ and $C^{\prime \prime}$ in a similar way.

Since the points $C$ and $B$ are the midpoints of arcs $T C^{\prime \prime}$ and $T B^{\prime \prime}$, respectively, we have

$$
\begin{aligned}
\angle\left(t, B^{\prime \prime} C^{\prime \prime}\right) & =\angle\left(t, T C^{\prime \prime}\right)+\angle\left(T C^{\prime \prime}, B^{\prime \prime} C^{\prime \prime}\right)=2 \angle(t, T C)+2 \angle\left(T C^{\prime \prime}, B C^{\prime \prime}\right) \\
& =2(\angle(t, T C)+\angle(T C, B C))=2 \angle(t, B C)=\angle\left(t, t_{a}\right) .
\end{aligned}
$$

It follows that $t_{a}$ and $B^{\prime \prime} C^{\prime \prime}$ are parallel. Similarly, $t_{b} \| A^{\prime \prime} C^{\prime \prime}$ and $t_{c} \| A^{\prime \prime} B^{\prime \prime}$. Thus, either the triangles $A^{\prime} B^{\prime} C^{\prime}$ and $A^{\prime \prime} B^{\prime \prime} C^{\prime \prime}$ are homothetic, or they are translates of each other. Now we will prove that they are in fact homothetic, and that the center $K$ of the homothety belongs
to $\omega$. It would then follow that their circumcircles are also homothetic with respect to $K$ and are therefore tangent at this point, as desired.

We need the two following claims.
Claim 1. The point of intersection $X$ of the lines $B^{\prime \prime} C$ and $B C^{\prime \prime}$ lies on $t_{a}$.
Proof. Actually, the points $X$ and $T$ are symmetric about the line $B C$, since the lines $C T$ and $C B^{\prime \prime}$ are symmetric about this line, as are the lines $B T$ and $B C^{\prime \prime}$.

Claim 2. The point of intersection $I$ of the lines $B B^{\prime}$ and $C C^{\prime}$ lies on the circle $\omega$.
Proof. We consider the case that $t$ is not parallel to the sides of $A B C$; the other cases may be regarded as limit cases. Let $D=t \cap B C, E=t \cap A C$, and $F=t \cap A B$.

Due to symmetry, the line $D B$ is one of the angle bisectors of the lines $B^{\prime} D$ and $F D$; analogously, the line $F B$ is one of the angle bisectors of the lines $B^{\prime} F$ and $D F$. So $B$ is either the incenter or one of the excenters of the triangle $B^{\prime} D F$. In any case we have $\angle(B D, D F)+\angle(D F, F B)+$ $\angle\left(B^{\prime} B, B^{\prime} D\right)=90^{\circ}$, so

$$
\angle\left(B^{\prime} B, B^{\prime} C^{\prime}\right)=\angle\left(B^{\prime} B, B^{\prime} D\right)=90^{\circ}-\angle(B C, D F)-\angle(D F, B A)=90^{\circ}-\angle(B C, A B) .
$$

Analogously, we get $\angle\left(C^{\prime} C, B^{\prime} C^{\prime}\right)=90^{\circ}-\angle(B C, A C)$. Hence,

$$
\angle(B I, C I)=\angle\left(B^{\prime} B, B^{\prime} C^{\prime}\right)+\angle\left(B^{\prime} C^{\prime}, C^{\prime} C\right)=\angle(B C, A C)-\angle(B C, A B)=\angle(A B, A C),
$$

which means exactly that the points $A, B, I, C$ are concyclic.
Now we can complete the proof. Let $K$ be the second intersection point of $B^{\prime} B^{\prime \prime}$ and $\omega$. Applying Pascal's theorem to hexagon $K B^{\prime \prime} C I B C^{\prime \prime}$ we get that the points $B^{\prime}=K B^{\prime \prime} \cap I B$ and $X=B^{\prime \prime} C \cap B C^{\prime \prime}$ are collinear with the intersection point $S$ of $C I$ and $C^{\prime \prime} K$. So $S=$ $C I \cap B^{\prime} X=C^{\prime}$, and the points $C^{\prime}, C^{\prime \prime}, K$ are collinear. Thus $K$ is the intersection point of $B^{\prime} B^{\prime \prime}$ and $C^{\prime} C^{\prime \prime}$ which implies that $K$ is the center of the homothety mapping $A^{\prime} B^{\prime} C^{\prime}$ to $A^{\prime \prime} B^{\prime \prime} C^{\prime \prime}$, and it belongs to $\omega$.

Solution 2. Define the points $T, A^{\prime}, B^{\prime}$, and $C^{\prime}$ in the same way as in the previous solution. Let $X, Y$, and $Z$ be the symmetric images of $T$ about the lines $B C, C A$, and $A B$, respectively. Note that the projections of $T$ on these lines form a Simson line of $T$ with respect to $A B C$, therefore the points $X, Y, Z$ are also collinear. Moreover, we have $X \in B^{\prime} C^{\prime}, Y \in C^{\prime} A^{\prime}$, $Z \in A^{\prime} B^{\prime}$.

Denote $\alpha=\angle(t, T C)=\angle(B T, B C)$. Using the symmetry in the lines $A C$ and $B C$, we get

$$
\angle(B C, B X)=\angle(B T, B C)=\alpha \quad \text { and } \quad \angle\left(X C, X C^{\prime}\right)=\angle(t, T C)=\angle\left(Y C, Y C^{\prime}\right)=\alpha .
$$

Since $\angle\left(X C, X C^{\prime}\right)=\angle\left(Y C, Y C^{\prime}\right)$, the points $X, Y, C, C^{\prime}$ lie on some circle $\omega_{c}$. Define the circles $\omega_{a}$ and $\omega_{b}$ analogously. Let $\omega^{\prime}$ be the circumcircle of triangle $A^{\prime} B^{\prime} C^{\prime}$.

Now, applying Miquel's theorem to the four lines $A^{\prime} B^{\prime}, A^{\prime} C^{\prime}, B^{\prime} C^{\prime}$, and $X Y$, we obtain that the circles $\omega^{\prime}, \omega_{a}, \omega_{b}, \omega_{c}$ intersect at some point $K$. We will show that $K$ lies on $\omega$, and that the tangent lines to $\omega$ and $\omega^{\prime}$ at this point coincide; this implies the problem statement.

Due to symmetry, we have $X B=T B=Z B$, so the point $B$ is the midpoint of one of the $\operatorname{arcs} X Z$ of circle $\omega_{b}$. Therefore $\angle(K B, K X)=\angle(X Z, X B)$. Analogously, $\angle(K X, K C)=$ $\angle(X C, X Y)$. Adding these equalities and using the symmetry in the line $B C$ we get

$$
\angle(K B, K C)=\angle(X Z, X B)+\angle(X C, X Z)=\angle(X C, X B)=\angle(T B, T C) .
$$

Therefore, $K$ lies on $\omega$.
Next, let $k$ be the tangent line to $\omega$ at $K$. We have

$$
\begin{aligned}
\angle\left(k, K C^{\prime}\right) & =\angle(k, K C)+\angle\left(K C, K C^{\prime}\right)=\angle(K B, B C)+\angle\left(X C, X C^{\prime}\right) \\
& =(\angle(K B, B X)-\angle(B C, B X))+\alpha=\angle\left(K B^{\prime}, B^{\prime} X\right)-\alpha+\alpha=\angle\left(K B^{\prime}, B^{\prime} C^{\prime}\right),
\end{aligned}
$$

which means exactly that $k$ is tangent to $\omega^{\prime}$.


Comment. There exist various solutions combining the ideas from the two solutions presented above. For instance, one may define the point $X$ as the reflection of $T$ with respect to the line $B C$, and then introduce the point $K$ as the second intersection point of the circumcircles of $B B^{\prime} X$ and $C C^{\prime} X$. Using the fact that $B B^{\prime}$ and $C C^{\prime}$ are the bisectors of $\angle\left(A^{\prime} B^{\prime}, B^{\prime} C^{\prime}\right)$ and $\angle\left(A^{\prime} C^{\prime}, B^{\prime} C^{\prime}\right)$ one can show successively that $K \in \omega, K \in \omega^{\prime}$, and that the tangents to $\omega$ and $\omega^{\prime}$ at $K$ coincide.

## N1

For any integer $d>0$, let $f(d)$ be the smallest positive integer that has exactly $d$ positive divisors (so for example we have $f(1)=1, f(5)=16$, and $f(6)=12$ ). Prove that for every integer $k \geq 0$ the number $f\left(2^{k}\right)$ divides $f\left(2^{k+1}\right)$.

Solution 1. For any positive integer $n$, let $d(n)$ be the number of positive divisors of $n$. Let $n=\prod_{p} p^{a(p)}$ be the prime factorization of $n$ where $p$ ranges over the prime numbers, the integers $a(p)$ are nonnegative and all but finitely many $a(p)$ are zero. Then we have $d(n)=\prod_{p}(a(p)+1)$. Thus, $d(n)$ is a power of 2 if and only if for every prime $p$ there is a nonnegative integer $b(p)$ with $a(p)=2^{b(p)}-1=1+2+2^{2}+\cdots+2^{b(p)-1}$. We then have

$$
n=\prod_{p} \prod_{i=0}^{b(p)-1} p^{2^{i}}, \quad \text { and } \quad d(n)=2^{k} \quad \text { with } \quad k=\sum_{p} b(p) .
$$

Let $\mathcal{S}$ be the set of all numbers of the form $p^{2^{r}}$ with $p$ prime and $r$ a nonnegative integer. Then we deduce that $d(n)$ is a power of 2 if and only if $n$ is the product of the elements of some finite subset $\mathcal{T}$ of $\mathcal{S}$ that satisfies the following condition: for all $t \in \mathcal{T}$ and $s \in \mathcal{S}$ with $s \mid t$ we have $s \in \mathcal{T}$. Moreover, if $d(n)=2^{k}$ then the corresponding set $\mathcal{T}$ has $k$ elements.

Note that the set $\mathcal{T}_{k}$ consisting of the smallest $k$ elements from $\mathcal{S}$ obviously satisfies the condition above. Thus, given $k$, the smallest $n$ with $d(n)=2^{k}$ is the product of the elements of $\mathcal{T}_{k}$. This $n$ is $f\left(2^{k}\right)$. Since obviously $\mathcal{T}_{k} \subset \mathcal{T}_{k+1}$, it follows that $f\left(2^{k}\right) \mid f\left(2^{k+1}\right)$.

Solution 2. This is an alternative to the second part of the Solution 1. Suppose $k$ is a nonnegative integer. From the first part of Solution 1 we see that $f\left(2^{k}\right)=\prod_{p} p^{a(p)}$ with $a(p)=2^{b(p)}-1$ and $\sum_{p} b(p)=k$. We now claim that for any two distinct primes $p, q$ with $b(q)>0$ we have

$$
\begin{equation*}
m=p^{2^{b(p)}}>q^{2^{b(q)-1}}=\ell . \tag{1}
\end{equation*}
$$

To see this, note first that $\ell$ divides $f\left(2^{k}\right)$. With the first part of Solution 1 one can see that the integer $n=f\left(2^{k}\right) m / \ell$ also satisfies $d(n)=2^{k}$. By the definition of $f\left(2^{k}\right)$ this implies that $n \geq f\left(2^{k}\right)$ so $m \geq \ell$. Since $p \neq q$ the inequality (1) follows.
Let the prime factorization of $f\left(2^{k+1}\right)$ be given by $f\left(2^{k+1}\right)=\prod_{p} p^{r(p)}$ with $r(p)=2^{s(p)}-1$. Since we have $\sum_{p} s(p)=k+1>k=\sum_{p} b(p)$ there is a prime $p$ with $s(p)>b(p)$. For any prime $q \neq p$ with $b(q)>0$ we apply inequality (1) twice and get

$$
q^{2^{s(q)}}>p^{2^{s(p)-1}} \geq p^{2^{b(p)}}>q^{2^{b(q)-1}}
$$

which implies $s(q) \geq b(q)$. It follows that $s(q) \geq b(q)$ for all primes $q$, so $f\left(2^{k}\right) \mid f\left(2^{k+1}\right)$.

## N2

Consider a polynomial $P(x)=\left(x+d_{1}\right)\left(x+d_{2}\right) \cdot \ldots \cdot\left(x+d_{9}\right)$, where $d_{1}, d_{2}, \ldots, d_{9}$ are nine distinct integers. Prove that there exists an integer $N$ such that for all integers $x \geq N$ the number $P(x)$ is divisible by a prime number greater than 20 .

Solution 1. Note that the statement of the problem is invariant under translations of $x$; hence without loss of generality we may suppose that the numbers $d_{1}, d_{2}, \ldots, d_{9}$ are positive.

The key observation is that there are only eight primes below 20 , while $P(x)$ involves more than eight factors.

We shall prove that $N=d^{8}$ satisfies the desired property, where $d=\max \left\{d_{1}, d_{2}, \ldots, d_{9}\right\}$. Suppose for the sake of contradiction that there is some integer $x \geq N$ such that $P(x)$ is composed of primes below 20 only. Then for every index $i \in\{1,2, \ldots, 9\}$ the number $x+d_{i}$ can be expressed as product of powers of the first 8 primes.

Since $x+d_{i}>x \geq d^{8}$ there is some prime power $f_{i}>d$ that divides $x+d_{i}$. Invoking the pigeonhole principle we see that there are two distinct indices $i$ and $j$ such that $f_{i}$ and $f_{j}$ are powers of the same prime number. For reasons of symmetry, we may suppose that $f_{i} \leq f_{j}$. Now both of the numbers $x+d_{i}$ and $x+d_{j}$ are divisible by $f_{i}$ and hence so is their difference $d_{i}-d_{j}$. But as

$$
0<\left|d_{i}-d_{j}\right| \leq \max \left(d_{i}, d_{j}\right) \leq d<f_{i}
$$

this is impossible. Thereby the problem is solved.

Solution 2. Observe that for each index $i \in\{1,2, \ldots, 9\}$ the product

$$
D_{i}=\prod_{1 \leq j \leq 9, j \neq i}\left|d_{i}-d_{j}\right|
$$

is positive. We claim that $N=\max \left\{D_{1}-d_{1}, D_{2}-d_{2}, \ldots, D_{9}-d_{9}\right\}+1$ satisfies the statement of the problem. Suppose there exists an integer $x \geq N$ such that all primes dividing $P(x)$ are smaller than 20. For each index $i$ we reduce the fraction $\left(x+d_{i}\right) / D_{i}$ to lowest terms. Since $x+d_{i}>D_{i}$ the numerator of the fraction we thereby get cannot be 1 , and hence it has to be divisible by some prime number $p_{i}<20$.

By the pigeonhole principle, there are a prime number $p$ and two distinct indices $i$ and $j$ such that $p_{i}=p_{j}=p$. Let $p^{\alpha_{i}}$ and $p^{\alpha_{j}}$ be the greatest powers of $p$ dividing $x+d_{i}$ and $x+d_{j}$, respectively. Due to symmetry we may suppose $\alpha_{i} \leq \alpha_{j}$. But now $p^{\alpha_{i}}$ divides $d_{i}-d_{j}$ and hence also $D_{i}$, which means that all occurrences of $p$ in the numerator of the fraction $\left(x+d_{i}\right) / D_{i}$ cancel out, contrary to the choice of $p=p_{i}$. This contradiction proves our claim.

Solution 3. Given a nonzero integer $N$ as well as a prime number $p$ we write $v_{p}(N)$ for the exponent with which $p$ occurs in the prime factorization of $|N|$.

Evidently, if the statement of the problem were not true, then there would exist an infinite sequence $\left(x_{n}\right)$ of positive integers tending to infinity such that for each $n \in \mathbb{Z}_{+}$the integer $P\left(x_{n}\right)$ is not divisible by any prime number $>20$. Observe that the numbers $-d_{1},-d_{2}, \ldots,-d_{9}$ do not appear in this sequence.

Now clearly there exists a prime $p_{1}<20$ for which the sequence $v_{p_{1}}\left(x_{n}+d_{1}\right)$ is not bounded; thinning out the sequence $\left(x_{n}\right)$ if necessary we may even suppose that

$$
v_{p_{1}}\left(x_{n}+d_{1}\right) \longrightarrow \infty .
$$

Repeating this argument eight more times we may similarly choose primes $p_{2}, \ldots, p_{9}<20$ and suppose that our sequence $\left(x_{n}\right)$ has been thinned out to such an extent that $v_{p_{i}}\left(x_{n}+d_{i}\right) \longrightarrow \infty$ holds for $i=2, \ldots, 9$ as well. In view of the pigeonhole principle, there are distinct indices $i$ and $j$ as well as a prime $p<20$ such that $p_{i}=p_{j}=p$. Setting $k=v_{p}\left(d_{i}-d_{j}\right)$ there now has to be some $n$ for which both $v_{p}\left(x_{n}+d_{i}\right)$ and $v_{p}\left(x_{n}+d_{j}\right)$ are greater than $k$. But now the numbers $x_{n}+d_{i}$ and $x_{n}+d_{j}$ are divisible by $p^{k+1}$ whilst their difference $d_{i}-d_{j}$ is not -a contradiction.
Comment. This problem is supposed to be a relatively easy one, so one might consider adding the hypothesis that the numbers $d_{1}, d_{2}, \ldots, d_{9}$ be positive. Then certain merely technical issues are not going to arise while the main ideas required to solve the problems remain the same.

Number Theory - solutions

## N3

Let $n \geq 1$ be an odd integer. Determine all functions $f$ from the set of integers to itself such that for all integers $x$ and $y$ the difference $f(x)-f(y)$ divides $x^{n}-y^{n}$.

Answer. All functions $f$ of the form $f(x)=\varepsilon x^{d}+c$, where $\varepsilon$ is in $\{1,-1\}$, the integer $d$ is a positive divisor of $n$, and $c$ is an integer.

Solution. Obviously, all functions in the answer satisfy the condition of the problem. We will show that there are no other functions satisfying that condition.

Let $f$ be a function satisfying the given condition. For each integer $n$, the function $g$ defined by $g(x)=f(x)+n$ also satisfies the same condition. Therefore, by subtracting $f(0)$ from $f(x)$ we may assume that $f(0)=0$.

For any prime $p$, the condition on $f$ with $(x, y)=(p, 0)$ states that $f(p)$ divides $p^{n}$. Since the set of primes is infinite, there exist integers $d$ and $\varepsilon$ with $0 \leq d \leq n$ and $\varepsilon \in\{1,-1\}$ such that for infinitely many primes $p$ we have $f(p)=\varepsilon p^{d}$. Denote the set of these primes by $P$. Since a function $g$ satisfies the given condition if and only if $-g$ satisfies the same condition, we may suppose $\varepsilon=1$.

The case $d=0$ is easily ruled out, because 0 does not divide any nonzero integer. Suppose $d \geq 1$ and write $n$ as $m d+r$, where $m$ and $r$ are integers such that $m \geq 1$ and $0 \leq r \leq d-1$. Let $x$ be an arbitrary integer. For each prime $p$ in $P$, the difference $f(p)-f(x)$ divides $p^{n}-x^{n}$. Using the equality $f(p)=p^{d}$, we get

$$
p^{n}-x^{n}=p^{r}\left(p^{d}\right)^{m}-x^{n} \equiv p^{r} f(x)^{m}-x^{n} \equiv 0 \quad\left(\bmod p^{d}-f(x)\right)
$$

Since we have $r<d$, for large enough primes $p \in P$ we obtain

$$
\left|p^{r} f(x)^{m}-x^{n}\right|<p^{d}-f(x)
$$

Hence $p^{r} f(x)^{m}-x^{n}$ has to be zero. This implies $r=0$ and $x^{n}=\left(x^{d}\right)^{m}=f(x)^{m}$. Since $m$ is odd, we obtain $f(x)=x^{d}$.

Comment. If $n$ is an even positive integer, then the functions $f$ of the form

$$
f(x)=\left\{\begin{array}{l}
x^{d}+c \text { for some integers }, \\
-x^{d}+c \text { for the rest of integers },
\end{array}\right.
$$

where $d$ is a positive divisor of $n / 2$ and $c$ is an integer, also satisfy the condition of the problem. Together with the functions in the answer, they are all functions that satisfy the condition when $n$ is even.

## N4

For each positive integer $k$, let $t(k)$ be the largest odd divisor of $k$. Determine all positive integers $a$ for which there exists a positive integer $n$ such that all the differences

$$
t(n+a)-t(n), \quad t(n+a+1)-t(n+1), \quad \ldots, \quad t(n+2 a-1)-t(n+a-1)
$$

are divisible by 4 .

Answer. $\quad a=1,3$, or 5 .

Solution. A pair $(a, n)$ satisfying the condition of the problem will be called a winning pair. It is straightforward to check that the pairs $(1,1),(3,1)$, and $(5,4)$ are winning pairs.

Now suppose that $a$ is a positive integer not equal to 1,3 , and 5 . We will show that there are no winning pairs ( $a, n$ ) by distinguishing three cases.

Case 1: $a$ is even. In this case we have $a=2^{\alpha} d$ for some positive integer $\alpha$ and some odd $d$. Since $a \geq 2^{\alpha}$, for each positive integer $n$ there exists an $i \in\{0,1, \ldots, a-1\}$ such that $n+i=2^{\alpha-1} e$, where $e$ is some odd integer. Then we have $t(n+i)=t\left(2^{\alpha-1} e\right)=e$ and

$$
t(n+a+i)=t\left(2^{\alpha} d+2^{\alpha-1} e\right)=2 d+e \equiv e+2 \quad(\bmod 4)
$$

So we get $t(n+i)-t(n+a+i) \equiv 2(\bmod 4)$, and $(a, n)$ is not a winning pair.
Case 2: $a$ is odd and $a>8$. For each positive integer $n$, there exists an $i \in\{0,1, \ldots, a-5\}$ such that $n+i=2 d$ for some odd $d$. We get

$$
t(n+i)=d \not \equiv d+2=t(n+i+4) \quad(\bmod 4)
$$

and

$$
t(n+a+i)=n+a+i \equiv n+a+i+4=t(n+a+i+4) \quad(\bmod 4) .
$$

Therefore, the integers $t(n+a+i)-t(n+i)$ and $t(n+a+i+4)-t(n+i+4)$ cannot be both divisible by 4 , and therefore there are no winning pairs in this case.

Case 3: $a=7$. For each positive integer $n$, there exists an $i \in\{0,1, \ldots, 6\}$ such that $n+i$ is either of the form $8 k+3$ or of the form $8 k+6$, where $k$ is a nonnegative integer. But we have

$$
t(8 k+3) \equiv 3 \not \equiv 1 \equiv 4 k+5=t(8 k+3+7) \quad(\bmod 4)
$$

and

$$
t(8 k+6)=4 k+3 \equiv 3 \not \equiv 1 \equiv t(8 k+6+7) \quad(\bmod 4) .
$$

Hence, there are no winning pairs of the form $(7, n)$.

## N5

Let $f$ be a function from the set of integers to the set of positive integers. Suppose that for any two integers $m$ and $n$, the difference $f(m)-f(n)$ is divisible by $f(m-n)$. Prove that for all integers $m, n$ with $f(m) \leq f(n)$ the number $f(n)$ is divisible by $f(m)$.

Solution 1. Suppose that $x$ and $y$ are two integers with $f(x)<f(y)$. We will show that $f(x) \mid f(y)$. By taking $m=x$ and $n=y$ we see that

$$
f(x-y)||f(x)-f(y)|=f(y)-f(x)>0
$$

so $f(x-y) \leq f(y)-f(x)<f(y)$. Hence the number $d=f(x)-f(x-y)$ satisfies

$$
-f(y)<-f(x-y)<d<f(x)<f(y)
$$

Taking $m=x$ and $n=x-y$ we see that $f(y) \mid d$, so we deduce $d=0$, or in other words $f(x)=f(x-y)$. Taking $m=x$ and $n=y$ we see that $f(x)=f(x-y) \mid f(x)-f(y)$, which implies $f(x) \mid f(y)$.

Solution 2. We split the solution into a sequence of claims; in each claim, the letters $m$ and $n$ denote arbitrary integers.

Claim 1. $f(n) \mid f(m n)$.
Proof. Since trivially $f(n) \mid f(1 \cdot n)$ and $f(n) \mid f((k+1) n)-f(k n)$ for all integers $k$, this is easily seen by using induction on $m$ in both directions.

Claim 2. $f(n) \mid f(0)$ and $f(n)=f(-n)$.
Proof. The first part follows by plugging $m=0$ into Claim 1. Using Claim 1 twice with $m=-1$, we get $f(n)|f(-n)| f(n)$, from which the second part follows.

From Claim 1, we get $f(1) \mid f(n)$ for all integers $n$, so $f(1)$ is the minimal value attained by $f$. Next, from Claim 2, the function $f$ can attain only a finite number of values since all these values divide $f(0)$.

Now we prove the statement of the problem by induction on the number $N_{f}$ of values attained by $f$. In the base case $N_{f} \leq 2$, we either have $f(0) \neq f(1)$, in which case these two numbers are the only values attained by $f$ and the statement is clear, or we have $f(0)=f(1)$, in which case we have $f(1)|f(n)| f(0)$ for all integers $n$, so $f$ is constant and the statement is obvious again.

For the induction step, assume that $N_{f} \geq 3$, and let $a$ be the least positive integer with $f(a)>f(1)$. Note that such a number exists due to the symmetry of $f$ obtained in Claim 2.

Claim 3. $f(n) \neq f(1)$ if and only if $a \mid n$.
Proof. Since $f(1)=\cdots=f(a-1)<f(a)$, the claim follows from the fact that

$$
f(n)=f(1) \Longleftrightarrow f(n+a)=f(1) .
$$

So it suffices to prove this fact.
Assume that $f(n)=f(1)$. Then $f(n+a) \mid f(a)-f(-n)=f(a)-f(n)>0$, so $f(n+a) \leq$ $f(a)-f(n)<f(a)$; in particular the difference $f(n+a)-f(n)$ is stricly smaller than $f(a)$. Furthermore, this difference is divisible by $f(a)$ and nonnegative since $f(n)=f(1)$ is the least value attained by $f$. So we have $f(n+a)-f(n)=0$, as desired. For the converse direction we only need to remark that $f(n+a)=f(1)$ entails $f(-n-a)=f(1)$, and hence $f(n)=f(-n)=f(1)$ by the forward implication.

We return to the induction step. So let us take two arbitrary integers $m$ and $n$ with $f(m) \leq f(n)$. If $a \nmid m$, then we have $f(m)=f(1) \mid f(n)$. On the other hand, suppose that $a \mid m$; then by Claim $3 a \mid n$ as well. Now define the function $g(x)=f(a x)$. Clearly, $g$ satisfies the conditions of the problem, but $N_{g}<N_{f}-1$, since $g$ does not attain $f(1)$. Hence, by the induction hypothesis, $f(m)=g(m / a) \mid g(n / a)=f(n)$, as desired.

Comment. After the fact that $f$ attains a finite number of values has been established, there are several ways of finishing the solution. For instance, let $f(0)=b_{1}>b_{2}>\cdots>b_{k}$ be all these values. One may show (essentially in the same way as in Claim 3) that the set $S_{i}=\left\{n: f(n) \geq b_{i}\right\}$ consists exactly of all numbers divisible by some integer $a_{i} \geq 0$. One obviously has $a_{i} \mid a_{i-1}$, which implies $f\left(a_{i}\right) \mid f\left(a_{i-1}\right)$ by Claim 1. So, $b_{k}\left|b_{k-1}\right| \cdots \mid b_{1}$, thus proving the problem statement.

Moreover, now it is easy to describe all functions satisfying the conditions of the problem. Namely, all these functions can be constructed as follows. Consider a sequence of nonnegative integers $a_{1}, a_{2}, \ldots, a_{k}$ and another sequence of positive integers $b_{1}, b_{2}, \ldots, b_{k}$ such that $\left|a_{k}\right|=1, a_{i} \neq a_{j}$ and $b_{i} \neq b_{j}$ for all $1 \leq i<j \leq k$, and $a_{i} \mid a_{i-1}$ and $b_{i} \mid b_{i-1}$ for all $i=2, \ldots, k$. Then one may introduce the function

$$
f(n)=b_{i(n)}, \quad \text { where } \quad i(n)=\min \left\{i: a_{i} \mid n\right\} .
$$

These are all the functions which satisfy the conditions of the problem.

## N6

Let $P(x)$ and $Q(x)$ be two polynomials with integer coefficients such that no nonconstant polynomial with rational coefficients divides both $P(x)$ and $Q(x)$. Suppose that for every positive integer $n$ the integers $P(n)$ and $Q(n)$ are positive, and $2^{Q(n)}-1$ divides $3^{P(n)}-1$. Prove that $Q(x)$ is a constant polynomial.

Solution. First we show that there exists an integer $d$ such that for all positive integers $n$ we have $\operatorname{gcd}(P(n), Q(n)) \leq d$.

Since $P(x)$ and $Q(x)$ are coprime (over the polynomials with rational coefficients), Euclid's algorithm provides some polynomials $R_{0}(x), S_{0}(x)$ with rational coefficients such that $P(x) R_{0}(x)-$ $Q(x) S_{0}(x)=1$. Multiplying by a suitable positive integer $d$, we obtain polynomials $R(x)=$ $d \cdot R_{0}(x)$ and $S(x)=d \cdot S_{0}(x)$ with integer coefficients for which $P(x) R(x)-Q(x) S(x)=d$. Then we have $\operatorname{gcd}(P(n), Q(n)) \leq d$ for any integer $n$.

To prove the problem statement, suppose that $Q(x)$ is not constant. Then the sequence $Q(n)$ is not bounded and we can choose a positive integer $m$ for which

$$
\begin{equation*}
M=2^{Q(m)}-1 \geq 3^{\max \{P(1), P(2), \ldots, P(d)\}} . \tag{1}
\end{equation*}
$$

Since $M=2^{Q(n)}-1 \mid 3^{P(n)}-1$, we have $2,3 \nmid M$. Let $a$ and $b$ be the multiplicative orders of 2 and 3 modulo $M$, respectively. Obviously, $a=Q(m)$ since the lower powers of 2 do not reach $M$. Since $M$ divides $3^{P(m)}-1$, we have $b \mid P(m)$. Then $\operatorname{gcd}(a, b) \leq \operatorname{gcd}(P(m), Q(m)) \leq d$. Since the expression $a x-b y$ attains all integer values divisible by $\operatorname{gcd}(a, b)$ when $x$ and $y$ run over all nonnegative integer values, there exist some nonnegative integers $x, y$ such that $1 \leq m+a x-b y \leq d$.

By $Q(m+a x) \equiv Q(m)(\bmod a)$ we have

$$
2^{Q(m+a x)} \equiv 2^{Q(m)} \equiv 1 \quad(\bmod M)
$$

and therefore

$$
M\left|2^{Q(m+a x)}-1\right| 3^{P(m+a x)}-1
$$

Then, by $P(m+a x-b y) \equiv P(m+a x)(\bmod b)$ we have

$$
3^{P(m+a x-b y)} \equiv 3^{P(m+a x)} \equiv 1 \quad(\bmod M)
$$

Since $P(m+a x-b y)>0$ this implies $M \leq 3^{P(m+a x-b y)}-1$. But $P(m+a x-b y)$ is listed among $P(1), P(2), \ldots, P(d)$, so

$$
M<3^{P(m+a x-b y)} \leq 3^{\max \{P(1), P(2), \ldots, P(d)\}}
$$

which contradicts (1).

Comment. We present another variant of the solution above.
Denote the degree of $P$ by $k$ and its leading coefficient by $p$. Consider any positive integer $n$ and let $a=Q(n)$. Again, denote by $b$ the multiplicative order of 3 modulo $2^{a}-1$. Since $2^{a}-1 \mid 3^{P(n)}-1$, we have $b \mid P(n)$. Moreover, since $2^{Q(n+a t)}-1 \mid 3^{P(n+a t)}-1$ and $a=Q(n) \mid Q(n+a t)$ for each positive integer $t$, we have $2^{a}-1 \mid 3^{P(n+a t)}-1$, hence $b \mid P(n+a t)$ as well.

Therefore, $b$ divides $\operatorname{gcd}\{P(n+a t): t \geq 0\}$; hence it also divides the number

$$
\sum_{i=0}^{k}(-1)^{k-i}\binom{k}{i} P(n+a i)=p \cdot k!\cdot a^{k} .
$$

Finally, we get $b \mid \operatorname{gcd}\left(P(n), k!\cdot p \cdot Q(n)^{k}\right)$, which is bounded by the same arguments as in the beginning of the solution. So $3^{b}-1$ is bounded, and hence $2^{Q(n)}-1$ is bounded as well.

N7
Let $p$ be an odd prime number. For every integer $a$, define the number

$$
S_{a}=\frac{a}{1}+\frac{a^{2}}{2}+\cdots+\frac{a^{p-1}}{p-1} .
$$

Let $m$ and $n$ be integers such that

$$
S_{3}+S_{4}-3 S_{2}=\frac{m}{n}
$$

Prove that $p$ divides $m$.

Solution 1. For rational numbers $p_{1} / q_{1}$ and $p_{2} / q_{2}$ with the denominators $q_{1}, q_{2}$ not divisible by $p$, we write $p_{1} / q_{1} \equiv p_{2} / q_{2}(\bmod p)$ if the numerator $p_{1} q_{2}-p_{2} q_{1}$ of their difference is divisible by $p$.

We start with finding an explicit formula for the residue of $S_{a}$ modulo $p$. Note first that for every $k=1, \ldots, p-1$ the number $\binom{p}{k}$ is divisible by $p$, and

$$
\frac{1}{p}\binom{p}{k}=\frac{(p-1)(p-2) \cdots(p-k+1)}{k!} \equiv \frac{(-1) \cdot(-2) \cdots(-k+1)}{k!}=\frac{(-1)^{k-1}}{k} \quad(\bmod p)
$$

Therefore, we have

$$
S_{a}=-\sum_{k=1}^{p-1} \frac{(-a)^{k}(-1)^{k-1}}{k} \equiv-\sum_{k=1}^{p-1}(-a)^{k} \cdot \frac{1}{p}\binom{p}{k} \quad(\bmod p) .
$$

The number on the right-hand side is integer. Using the binomial formula we express it as

$$
-\sum_{k=1}^{p-1}(-a)^{k} \cdot \frac{1}{p}\binom{p}{k}=-\frac{1}{p}\left(-1-(-a)^{p}+\sum_{k=0}^{p}(-a)^{k}\binom{p}{k}\right)=\frac{(a-1)^{p}-a^{p}+1}{p}
$$

since $p$ is odd. So, we have

$$
S_{a} \equiv \frac{(a-1)^{p}-a^{p}+1}{p} \quad(\bmod p)
$$

Finally, using the obtained formula we get

$$
\begin{aligned}
S_{3}+S_{4}-3 S_{2} & \equiv \frac{\left(2^{p}-3^{p}+1\right)+\left(3^{p}-4^{p}+1\right)-3\left(1^{p}-2^{p}+1\right)}{p} \\
& =\frac{4 \cdot 2^{p}-4^{p}-4}{p}=-\frac{\left(2^{p}-2\right)^{2}}{p} \quad(\bmod p) .
\end{aligned}
$$

By Fermat's theorem, $p \mid 2^{p}-2$, so $p^{2} \mid\left(2^{p}-2\right)^{2}$ and hence $S_{3}+S_{4}-3 S_{2} \equiv 0(\bmod p)$.

Solution 2. One may solve the problem without finding an explicit formula for $S_{a}$. It is enough to find the following property.

Lemma. For every integer $a$, we have $S_{a+1} \equiv S_{-a}(\bmod p)$.
Proof. We expand $S_{a+1}$ using the binomial formula as

$$
S_{a+1}=\sum_{k=1}^{p-1} \frac{1}{k} \sum_{j=0}^{k}\binom{k}{j} a^{j}=\sum_{k=1}^{p-1}\left(\frac{1}{k}+\sum_{j=1}^{k} a^{j} \cdot \frac{1}{k}\binom{k}{j}\right)=\sum_{k=1}^{p-1} \frac{1}{k}+\sum_{j=1}^{p-1} a^{j} \sum_{k=j}^{p-1} \frac{1}{k}\binom{k}{j} a^{k} .
$$

Note that $\frac{1}{k}+\frac{1}{p-k}=\frac{p}{k(p-k)} \equiv 0(\bmod p)$ for all $1 \leq k \leq p-1$; hence the first sum vanishes modulo $p$. For the second sum, we use the relation $\frac{1}{k}\binom{k}{j}=\frac{1}{j}\binom{k-1}{j-1}$ to obtain

$$
S_{a+1} \equiv \sum_{j=1}^{p-1} \frac{a^{j}}{j} \sum_{k=1}^{p-1}\binom{k-1}{j-1} \quad(\bmod p) .
$$

Finally, from the relation

$$
\sum_{k=1}^{p-1}\binom{k-1}{j-1}=\binom{p-1}{j}=\frac{(p-1)(p-2) \ldots(p-j)}{j!} \equiv(-1)^{j} \quad(\bmod p)
$$

we obtain

$$
S_{a+1} \equiv \sum_{j=1}^{p-1} \frac{a^{j}(-1)^{j}}{j!}=S_{-a} .
$$

Now we turn to the problem. Using the lemma we get

$$
\begin{equation*}
S_{3}-3 S_{2} \equiv S_{-2}-3 S_{2}=\sum_{\substack{1 \leq k \leq p-1 \\ k \text { is even }}} \frac{-2 \cdot 2^{k}}{k}+\sum_{\substack{1 \leq k \leq p-1 \\ k \text { is odd }}} \frac{-4 \cdot 2^{k}}{k}(\bmod p) \tag{1}
\end{equation*}
$$

The first sum in (1) expands as

$$
\sum_{\ell=1}^{(p-1) / 2} \frac{-2 \cdot 2^{2 \ell}}{2 \ell}=-\sum_{\ell=1}^{(p-1) / 2} \frac{4^{\ell}}{\ell}
$$

Next, using Fermat's theorem, we expand the second sum in (11) as

$$
-\sum_{\ell=1}^{(p-1) / 2} \frac{2^{2 \ell+1}}{2 \ell-1} \equiv-\sum_{\ell=1}^{(p-1) / 2} \frac{2^{p+2 \ell}}{p+2 \ell-1}=-\sum_{m=(p+1) / 2}^{p-1} \frac{2 \cdot 4^{m}}{2 m}=-\sum_{m=(p+1) / 2}^{p-1} \frac{4^{m}}{m} \quad(\bmod p)
$$

(here we set $m=\ell+\frac{p-1}{2}$ ). Hence,

$$
S_{3}-3 S_{2} \equiv-\sum_{\ell=1}^{(p-1) / 2} \frac{4^{\ell}}{\ell}-\sum_{m=(p+1) / 2}^{p-1} \frac{4^{m}}{m}=-S_{4} \quad(\bmod p) .
$$

## N8

Let $k$ be a positive integer and set $n=2^{k}+1$. Prove that $n$ is a prime number if and only if the following holds: there is a permutation $a_{1}, \ldots, a_{n-1}$ of the numbers $1,2, \ldots, n-1$ and a sequence of integers $g_{1}, g_{2}, \ldots, g_{n-1}$ such that $n$ divides $g_{i}^{a_{i}}-a_{i+1}$ for every $i \in\{1,2, \ldots, n-1\}$, where we set $a_{n}=a_{1}$.

Solution. Let $N=\{1,2, \ldots, n-1\}$. For $a, b \in N$, we say that $b$ follows $a$ if there exists an integer $g$ such that $b \equiv g^{a}(\bmod n)$ and denote this property as $a \rightarrow b$. This way we have a directed graph with $N$ as set of vertices. If $a_{1}, \ldots, a_{n-1}$ is a permutation of $1,2, \ldots, n-1$ such that $a_{1} \rightarrow a_{2} \rightarrow \ldots \rightarrow a_{n-1} \rightarrow a_{1}$ then this is a Hamiltonian cycle in the graph.

Step I. First consider the case when $n$ is composite. Let $n=p_{1}^{\alpha_{1}} \ldots p_{s}^{\alpha_{s}}$ be its prime factorization. All primes $p_{i}$ are odd.

Suppose that $\alpha_{i}>1$ for some $i$. For all integers $a, g$ with $a \geq 2$, we have $g^{a} \not \equiv p_{i}\left(\bmod p_{i}^{2}\right)$, because $g^{a}$ is either divisible by $p_{i}^{2}$ or it is not divisible by $p_{i}$. It follows that in any Hamiltonian cycle $p_{i}$ comes immediately after 1 . The same argument shows that $2 p_{i}$ also should come immediately after 1 , which is impossible. Hence, there is no Hamiltonian cycle in the graph.
Now suppose that $n$ is square-free. We have $n=p_{1} p_{2} \ldots p_{s}>9$ and $s \geq 2$. Assume that there exists a Hamiltonian cycle. There are $\frac{n-1}{2}$ even numbers in this cycle, and each number which follows one of them should be a quadratic residue modulo $n$. So, there should be at least $\frac{n-1}{2}$ nonzero quadratic residues modulo $n$. On the other hand, for each $p_{i}$ there exist exactly $\frac{p_{i}+1}{2}$ quadratic residues modulo $p_{i}$; by the Chinese Remainder Theorem, the number of quadratic residues modulo $n$ is exactly $\frac{p_{1}+1}{2} \cdot \frac{p_{2}+1}{2} \cdot \ldots \cdot \frac{p_{s}+1}{2}$, including 0 . Then we have a contradiction by

$$
\frac{p_{1}+1}{2} \cdot \frac{p_{2}+1}{2} \cdot \ldots \cdot \frac{p_{s}+1}{2} \leq \frac{2 p_{1}}{3} \cdot \frac{2 p_{2}}{3} \cdot \ldots \cdot \frac{2 p_{s}}{3}=\left(\frac{2}{3}\right)^{s} n \leq \frac{4 n}{9}<\frac{n-1}{2}
$$

This proves the "if"-part of the problem.
Step II. Now suppose that $n$ is prime. For any $a \in N$, denote by $\nu_{2}(a)$ the exponent of 2 in the prime factorization of $a$, and let $\mu(a)=\max \left\{t \in[0, k] \mid 2^{t} \rightarrow a\right\}$.

Lemma. For any $a, b \in N$, we have $a \rightarrow b$ if and only if $\nu_{2}(a) \leq \mu(b)$.
Proof. Let $\ell=\nu_{2}(a)$ and $m=\mu(b)$.
Suppose $\ell \leq m$. Since $b$ follows $2^{m}$, there exists some $g_{0}$ such that $b \equiv g_{0}^{2^{m}}(\bmod n)$. By $\operatorname{gcd}(a, n-1)=2^{\ell}$ there exist some integers $p$ and $q$ such that $p a-q(n-1)=2^{\ell}$. Choosing $g=g_{0}^{2^{m-\ell} p}$ we have $g^{a}=g_{0}^{2^{m-\ell} p a}=g_{0}^{2^{m}+2^{m-\ell} q(n-1)} \equiv g_{0}^{2^{m}} \equiv b(\bmod n)$ by Fermat's theorem. Hence, $a \rightarrow b$.

To prove the reverse statement, suppose that $a \rightarrow b$, so $b \equiv g^{a}(\bmod n)$ with some $g$. Then $b \equiv\left(g^{a / 2^{\ell}}\right)^{2^{\ell}}$, and therefore $2^{\ell} \rightarrow b$. By the definition of $\mu(b)$, we have $\mu(b) \geq \ell$. The lemma is
proved.
Now for every $i$ with $0 \leq i \leq k$, let

$$
\begin{aligned}
A_{i} & =\left\{a \in N \mid \nu_{2}(a)=i\right\}, \\
B_{i} & =\{a \in N \mid \mu(a)=i\}, \\
\text { and } C_{i} & =\{a \in N \mid \mu(a) \geq i\}=B_{i} \cup B_{i+1} \cup \ldots \cup B_{k} .
\end{aligned}
$$

We claim that $\left|A_{i}\right|=\left|B_{i}\right|$ for all $0 \leq i \leq k$. Obviously we have $\left|A_{i}\right|=2^{k-i-1}$ for all $i=$ $0, \ldots, k-1$, and $\left|A_{k}\right|=1$. Now we determine $\left|C_{i}\right|$. We have $\left|C_{0}\right|=n-1$ and by Fermat's theorem we also have $C_{k}=\{1\}$, so $\left|C_{k}\right|=1$. Next, notice that $C_{i+1}=\left\{x^{2} \bmod n \mid x \in C_{i}\right\}$. For every $a \in N$, the relation $x^{2} \equiv a(\bmod n)$ has at most two solutions in $N$. Therefore we have $2\left|C_{i+1}\right| \leq\left|C_{i}\right|$, with the equality achieved only if for every $y \in C_{i+1}$, there exist distinct elements $x, x^{\prime} \in C_{i}$ such that $x^{2} \equiv x^{\prime 2} \equiv y(\bmod n)$ (this implies $x+x^{\prime}=n$ ). Now, since $2^{k}\left|C_{k}\right|=\left|C_{0}\right|$, we obtain that this equality should be achieved in each step. Hence $\left|C_{i}\right|=2^{k-i}$ for $0 \leq i \leq k$, and therefore $\left|B_{i}\right|=2^{k-i-1}$ for $0 \leq i \leq k-1$ and $\left|B_{k}\right|=1$.

From the previous arguments we can see that for each $z \in C_{i}(0 \leq i<k)$ the equation $x^{2} \equiv z^{2}$ $(\bmod n)$ has two solutions in $C_{i}$, so we have $n-z \in C_{i}$. Hence, for each $i=0,1, \ldots, k-1$, exactly half of the elements of $C_{i}$ are odd. The same statement is valid for $B_{i}=C_{i} \backslash C_{i+1}$ for $0 \leq i \leq k-2$. In particular, each such $B_{i}$ contains an odd number. Note that $B_{k}=\{1\}$ also contains an odd number, and $B_{k-1}=\left\{2^{k}\right\}$ since $C_{k-1}$ consists of the two square roots of 1 modulo $n$.

Step III. Now we construct a Hamiltonian cycle in the graph. First, for each $i$ with $0 \leq i \leq k$, connect the elements of $A_{i}$ to the elements of $B_{i}$ by means of an arbitrary bijection. After performing this for every $i$, we obtain a subgraph with all vertices having in-degree 1 and outdegree 1 , so the subgraph is a disjoint union of cycles. If there is a unique cycle, we are done. Otherwise, we modify the subgraph in such a way that the previous property is preserved and the number of cycles decreases; after a finite number of steps we arrive at a single cycle.

For every cycle $C$, let $\lambda(C)=\min _{c \in C} \nu_{2}(c)$. Consider a cycle $C$ for which $\lambda(C)$ is maximal. If $\lambda(C)=0$, then for any other cycle $C^{\prime}$ we have $\lambda\left(C^{\prime}\right)=0$. Take two arbitrary vertices $a \in C$ and $a^{\prime} \in C^{\prime}$ such that $\nu_{2}(a)=\nu_{2}\left(a^{\prime}\right)=0$; let their direct successors be $b$ and $b^{\prime}$, respectively. Then we can unify $C$ and $C^{\prime}$ to a single cycle by replacing the edges $a \rightarrow b$ and $a^{\prime} \rightarrow b^{\prime}$ by $a \rightarrow b^{\prime}$ and $a^{\prime} \rightarrow b$.

Now suppose that $\lambda=\lambda(C) \geq 1$; let $a \in C \cap A_{\lambda}$. If there exists some $a^{\prime} \in A_{\lambda} \backslash C$, then $a^{\prime}$ lies in another cycle $C^{\prime}$ and we can merge the two cycles in exactly the same way as above. So, the only remaining case is $A_{\lambda} \subset C$. Since the edges from $A_{\lambda}$ lead to $B_{\lambda}$, we get also $B_{\lambda} \subset C$. If $\lambda \neq k-1$ then $B_{\lambda}$ contains an odd number; this contradicts the assumption $\lambda(C)>0$. Finally, if $\lambda=k-1$, then $C$ contains $2^{k-1}$ which is the only element of $A_{k-1}$. Since $B_{k-1}=\left\{2^{k}\right\}=A_{k}$ and $B_{k}=\{1\}$, the cycle $C$ contains the path $2^{k-1} \rightarrow 2^{k} \rightarrow 1$ and it contains an odd number again. This completes the proof of the "only if"-part of the problem.

Comment 1. The lemma and the fact $\left|A_{i}\right|=\left|B_{i}\right|$ together show that for every edge $a \rightarrow b$ of the Hamiltonian cycle, $\nu_{2}(a)=\mu(b)$ must hold. After this observation, the Hamiltonian cycle can be built in many ways. For instance, it is possible to select edges from $A_{i}$ to $B_{i}$ for $i=k, k-1, \ldots, 1$ in such a way that they form disjoint paths; at the end all these paths will have odd endpoints. In the final step, the paths can be closed to form a unique cycle.

Comment 2. Step II is an easy consequence of some basic facts about the multiplicative group modulo the prime $n=2^{k}+1$. The Lemma follows by noting that this group has order $2^{k}$, so the $a$-th powers are exactly the $2^{\nu_{2}(a)}$-th powers. Using the existence of a primitive root $g$ modulo $n$ one sees that the map from $\{1,2, \ldots, n-1\}$ to itself that sends $a$ to $g^{a} \bmod n$ is a bijection that sends $A_{i}$ to $B_{i}$ for each $i \in\{0, \ldots, k\}$.

# Shortlisted Problems with Solutions 

$53^{\text {rd }}$ International Mathematical Olympiad Mar del Plata, Argentina 2012

# The shortlisted problems should be kept strictly confidential until IMO 2013 

## Contributing Countries

The Organizing Committee and the Problem Selection Committee of IMO 2012 thank the following 40 countries for contributing 136 problem proposals:

Australia, Austria, Belarus, Belgium, Bulgaria, Canada, Cyprus, Czech Republic, Denmark, Estonia, Finland, France, Germany, Greece, Hong Kong, India, Iran, Ireland, Israel, Japan, Kazakhstan, Luxembourg, Malaysia, Montenegro, Netherlands, Norway, Pakistan, Romania, Russia, Serbia, Slovakia, Slovenia, South Africa, South Korea, Sweden, Thailand, Ukraine, United Kingdom, United States of America, Uzbekistan

## Problem Selection Committee

Martín Avendaño
Carlos di Fiore
Géza Kós
Svetoslav Savchev

## Algebra

A1. Find all the functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that

$$
f(a)^{2}+f(b)^{2}+f(c)^{2}=2 f(a) f(b)+2 f(b) f(c)+2 f(c) f(a)
$$

for all integers $a, b, c$ satisfying $a+b+c=0$.
A2. Let $\mathbb{Z}$ and $\mathbb{Q}$ be the sets of integers and rationals respectively.
a) Does there exist a partition of $\mathbb{Z}$ into three non-empty subsets $A, B, C$ such that the sets $A+B, B+C, C+A$ are disjoint?
b) Does there exist a partition of $\mathbb{Q}$ into three non-empty subsets $A, B, C$ such that the sets $A+B, B+C, C+A$ are disjoint?

Here $X+Y$ denotes the set $\{x+y \mid x \in X, y \in Y\}$, for $X, Y \subseteq \mathbb{Z}$ and $X, Y \subseteq \mathbb{Q}$.
A3. Let $a_{2}, \ldots, a_{n}$ be $n-1$ positive real numbers, where $n \geq 3$, such that $a_{2} a_{3} \cdots a_{n}=1$. Prove that

$$
\left(1+a_{2}\right)^{2}\left(1+a_{3}\right)^{3} \cdots\left(1+a_{n}\right)^{n}>n^{n} .
$$

A4. Let $f$ and $g$ be two nonzero polynomials with integer coefficients and $\operatorname{deg} f>\operatorname{deg} g$. Suppose that for infinitely many primes $p$ the polynomial $p f+g$ has a rational root. Prove that $f$ has a rational root.

A5. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ that satisfy the conditions

$$
f(1+x y)-f(x+y)=f(x) f(y) \quad \text { for all } x, y \in \mathbb{R}
$$

and $f(-1) \neq 0$.
A6. Let $f: \mathbb{N} \rightarrow \mathbb{N}$ be a function, and let $f^{m}$ be $f$ applied $m$ times. Suppose that for every $n \in \mathbb{N}$ there exists a $k \in \mathbb{N}$ such that $f^{2 k}(n)=n+k$, and let $k_{n}$ be the smallest such $k$. Prove that the sequence $k_{1}, k_{2}, \ldots$ is unbounded.

A7. We say that a function $f: \mathbb{R}^{k} \rightarrow \mathbb{R}$ is a metapolynomial if, for some positive integers $m$ and $n$, it can be represented in the form

$$
f\left(x_{1}, \ldots, x_{k}\right)=\max _{i=1, \ldots, m} \min _{j=1, \ldots, n} P_{i, j}\left(x_{1}, \ldots, x_{k}\right)
$$

where $P_{i, j}$ are multivariate polynomials. Prove that the product of two metapolynomials is also a metapolynomial.

## Combinatorics

C1. Several positive integers are written in a row. Iteratively, Alice chooses two adjacent numbers $x$ and $y$ such that $x>y$ and $x$ is to the left of $y$, and replaces the pair $(x, y)$ by either $(y+1, x)$ or $(x-1, x)$. Prove that she can perform only finitely many such iterations.

C2. Let $n \geq 1$ be an integer. What is the maximum number of disjoint pairs of elements of the set $\{1,2, \ldots, n\}$ such that the sums of the different pairs are different integers not exceeding $n$ ?

C3. In a $999 \times 999$ square table some cells are white and the remaining ones are red. Let $T$ be the number of triples $\left(C_{1}, C_{2}, C_{3}\right)$ of cells, the first two in the same row and the last two in the same column, with $C_{1}$ and $C_{3}$ white and $C_{2}$ red. Find the maximum value $T$ can attain.

C4. Players $A$ and $B$ play a game with $N \geq 2012$ coins and 2012 boxes arranged around a circle. Initially $A$ distributes the coins among the boxes so that there is at least 1 coin in each box. Then the two of them make moves in the order $B, A, B, A, \ldots$ by the following rules:

- On every move of his $B$ passes 1 coin from every box to an adjacent box.
- On every move of hers $A$ chooses several coins that were not involved in $B$ 's previous move and are in different boxes. She passes every chosen coin to an adjacent box.

Player $A$ 's goal is to ensure at least 1 coin in each box after every move of hers, regardless of how $B$ plays and how many moves are made. Find the least $N$ that enables her to succeed.

C5. The columns and the rows of a $3 n \times 3 n$ square board are numbered $1,2, \ldots, 3 n$. Every square $(x, y)$ with $1 \leq x, y \leq 3 n$ is colored asparagus, byzantium or citrine according as the modulo 3 remainder of $x+y$ is 0,1 or 2 respectively. One token colored asparagus, byzantium or citrine is placed on each square, so that there are $3 n^{2}$ tokens of each color.

Suppose that one can permute the tokens so that each token is moved to a distance of at most $d$ from its original position, each asparagus token replaces a byzantium token, each byzantium token replaces a citrine token, and each citrine token replaces an asparagus token. Prove that it is possible to permute the tokens so that each token is moved to a distance of at most $d+2$ from its original position, and each square contains a token with the same color as the square.

C6. Let $k$ and $n$ be fixed positive integers. In the liar's guessing game, Amy chooses integers $x$ and $N$ with $1 \leq x \leq N$. She tells Ben what $N$ is, but not what $x$ is. Ben may then repeatedly ask Amy whether $x \in S$ for arbitrary sets $S$ of integers. Amy will always answer with yes or no, but she might lie. The only restriction is that she can lie at most $k$ times in a row. After he has asked as many questions as he wants, Ben must specify a set of at most $n$ positive integers. If $x$ is in this set he wins; otherwise, he loses. Prove that:
a) If $n \geq 2^{k}$ then Ben can always win.
b) For sufficiently large $k$ there exist $n \geq 1.99^{k}$ such that Ben cannot guarantee a win.

C7. There are given $2^{500}$ points on a circle labeled $1,2, \ldots, 2^{500}$ in some order. Prove that one can choose 100 pairwise disjoint chords joining some of these points so that the 100 sums of the pairs of numbers at the endpoints of the chosen chords are equal.

## Geometry

G1. In the triangle $A B C$ the point $J$ is the center of the excircle opposite to $A$. This excircle is tangent to the side $B C$ at $M$, and to the lines $A B$ and $A C$ at $K$ and $L$ respectively. The lines $L M$ and $B J$ meet at $F$, and the lines $K M$ and $C J$ meet at $G$. Let $S$ be the point of intersection of the lines $A F$ and $B C$, and let $T$ be the point of intersection of the lines $A G$ and $B C$. Prove that $M$ is the midpoint of $S T$.

G2. Let $A B C D$ be a cyclic quadrilateral whose diagonals $A C$ and $B D$ meet at $E$. The extensions of the sides $A D$ and $B C$ beyond $A$ and $B$ meet at $F$. Let $G$ be the point such that $E C G D$ is a parallelogram, and let $H$ be the image of $E$ under reflection in $A D$. Prove that $D, H, F, G$ are concyclic.

G3. In an acute triangle $A B C$ the points $D, E$ and $F$ are the feet of the altitudes through $A$, $B$ and $C$ respectively. The incenters of the triangles $A E F$ and $B D F$ are $I_{1}$ and $I_{2}$ respectively; the circumcenters of the triangles $A C I_{1}$ and $B C I_{2}$ are $O_{1}$ and $O_{2}$ respectively. Prove that $I_{1} I_{2}$ and $O_{1} O_{2}$ are parallel.

G4. Let $A B C$ be a triangle with $A B \neq A C$ and circumcenter $O$. The bisector of $\angle B A C$ intersects $B C$ at $D$. Let $E$ be the reflection of $D$ with respect to the midpoint of $B C$. The lines through $D$ and $E$ perpendicular to $B C$ intersect the lines $A O$ and $A D$ at $X$ and $Y$ respectively. Prove that the quadrilateral $B X C Y$ is cyclic.

G5. Let $A B C$ be a triangle with $\angle B C A=90^{\circ}$, and let $C_{0}$ be the foot of the altitude from $C$. Choose a point $X$ in the interior of the segment $C C_{0}$, and let $K, L$ be the points on the segments $A X, B X$ for which $B K=B C$ and $A L=A C$ respectively. Denote by $M$ the intersection of $A L$ and $B K$. Show that $M K=M L$.

G6. Let $A B C$ be a triangle with circumcenter $O$ and incenter $I$. The points $D, E$ and $F$ on the sides $B C, C A$ and $A B$ respectively are such that $B D+B F=C A$ and $C D+C E=A B$. The circumcircles of the triangles $B F D$ and $C D E$ intersect at $P \neq D$. Prove that $O P=O I$.

G7. Let $A B C D$ be a convex quadrilateral with non-parallel sides $B C$ and $A D$. Assume that there is a point $E$ on the side $B C$ such that the quadrilaterals $A B E D$ and $A E C D$ are circumscribed. Prove that there is a point $F$ on the side $A D$ such that the quadrilaterals $A B C F$ and $B C D F$ are circumscribed if and only if $A B$ is parallel to $C D$.

G8. Let $A B C$ be a triangle with circumcircle $\omega$ and $\ell$ a line without common points with $\omega$. Denote by $P$ the foot of the perpendicular from the center of $\omega$ to $\ell$. The side-lines $B C, C A, A B$ intersect $\ell$ at the points $X, Y, Z$ different from $P$. Prove that the circumcircles of the triangles $A X P, B Y P$ and $C Z P$ have a common point different from $P$ or are mutually tangent at $P$.

## Number Theory

N1. Call admissible a set $A$ of integers that has the following property:

$$
\text { If } x, y \in A \text { (possibly } x=y \text { ) then } x^{2}+k x y+y^{2} \in A \text { for every integer } k \text {. }
$$

Determine all pairs $m, n$ of nonzero integers such that the only admissible set containing both $m$ and $n$ is the set of all integers.

N2. Find all triples $(x, y, z)$ of positive integers such that $x \leq y \leq z$ and

$$
x^{3}\left(y^{3}+z^{3}\right)=2012(x y z+2) .
$$

N3. Determine all integers $m \geq 2$ such that every $n$ with $\frac{m}{3} \leq n \leq \frac{m}{2}$ divides the binomial coefficient $\binom{n}{m-2 n}$.

N4. An integer $a$ is called friendly if the equation $\left(m^{2}+n\right)\left(n^{2}+m\right)=a(m-n)^{3}$ has a solution over the positive integers.
a) Prove that there are at least 500 friendly integers in the set $\{1,2, \ldots, 2012\}$.
b) Decide whether $a=2$ is friendly.

N5. For a nonnegative integer $n$ define $\operatorname{rad}(n)=1$ if $n=0$ or $n=1$, and $\operatorname{rad}(n)=p_{1} p_{2} \cdots p_{k}$ where $p_{1}<p_{2}<\cdots<p_{k}$ are all prime factors of $n$. Find all polynomials $f(x)$ with nonnegative integer coefficients such that $\operatorname{rad}(f(n))$ divides $\operatorname{rad}\left(f\left(n^{\operatorname{rad}(n)}\right)\right)$ for every nonnegative integer $n$.

N6. Let $x$ and $y$ be positive integers. If $x^{2^{n}}-1$ is divisible by $2^{n} y+1$ for every positive integer $n$, prove that $x=1$.

N7. Find all $n \in \mathbb{N}$ for which there exist nonnegative integers $a_{1}, a_{2}, \ldots, a_{n}$ such that

$$
\frac{1}{2^{a_{1}}}+\frac{1}{2^{a_{2}}}+\cdots+\frac{1}{2^{a_{n}}}=\frac{1}{3^{a_{1}}}+\frac{2}{3^{a_{2}}}+\cdots+\frac{n}{3^{a_{n}}}=1 .
$$

N8. Prove that for every prime $p>100$ and every integer $r$ there exist two integers $a$ and $b$ such that $p$ divides $a^{2}+b^{5}-r$.

## Algebra

A1. Find all the functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that

$$
f(a)^{2}+f(b)^{2}+f(c)^{2}=2 f(a) f(b)+2 f(b) f(c)+2 f(c) f(a)
$$

for all integers $a, b, c$ satisfying $a+b+c=0$.
Solution. The substitution $a=b=c=0$ gives $3 f(0)^{2}=6 f(0)^{2}$, hence

$$
\begin{equation*}
f(0)=0 \tag{1}
\end{equation*}
$$

The substitution $b=-a$ and $c=0$ gives $\left((f(a)-f(-a))^{2}=0\right.$. Hence $f$ is an even function:

$$
\begin{equation*}
f(a)=f(-a) \quad \text { for all } a \in \mathbb{Z} \tag{2}
\end{equation*}
$$

Now set $b=a$ and $c=-2 a$ to obtain $2 f(a)^{2}+f(2 a)^{2}=2 f(a)^{2}+4 f(a) f(2 a)$. Hence

$$
\begin{equation*}
f(2 a)=0 \text { or } f(2 a)=4 f(a) \quad \text { for all } a \in \mathbb{Z} \tag{3}
\end{equation*}
$$

If $f(r)=0$ for some $r \geq 1$ then the substitution $b=r$ and $c=-a-r$ gives $(f(a+r)-f(a))^{2}=0$. So $f$ is periodic with period $r$, i. e.

$$
f(a+r)=f(a) \quad \text { for all } a \in \mathbb{Z}
$$

In particular, if $f(1)=0$ then $f$ is constant, thus $f(a)=0$ for all $a \in \mathbb{Z}$. This function clearly satisfies the functional equation. For the rest of the analysis, we assume $f(1)=k \neq 0$.

By (3) we have $f(2)=0$ or $f(2)=4 k$. If $f(2)=0$ then $f$ is periodic of period 2 , thus $f($ even $)=0$ and $f($ odd $)=k$. This function is a solution for every $k$. We postpone the verification; for the sequel assume $f(2)=4 k \neq 0$.

By (3) again, we have $f(4)=0$ or $f(4)=16 k$. In the first case $f$ is periodic of period 4 , and $f(3)=f(-1)=f(1)=k$, so we have $f(4 n)=0, f(4 n+1)=f(4 n+3)=k$, and $f(4 n+2)=4 k$ for all $n \in \mathbb{Z}$. This function is a solution too, which we justify later. For the rest of the analysis, we assume $f(4)=16 k \neq 0$.

We show now that $f(3)=9 k$. In order to do so, we need two substitutions:

$$
\begin{gathered}
a=1, b=2, c=-3 \Longrightarrow f(3)^{2}-10 k f(3)+9 k^{2}=0 \Longrightarrow f(3) \in\{k, 9 k\} \\
a=1, b=3, c=-4 \Longrightarrow f(3)^{2}-34 k f(3)+225 k^{2}=0 \Longrightarrow f(3) \in\{9 k, 25 k\}
\end{gathered}
$$

Therefore $f(3)=9 k$, as claimed. Now we prove inductively that the only remaining function is $f(x)=k x^{2}, x \in \mathbb{Z}$. We proved this for $x=0,1,2,3,4$. Assume that $n \geq 4$ and that $f(x)=k x^{2}$ holds for all integers $x \in[0, n]$. Then the substitutions $a=n, b=1, c=-n-1$ and $a=n-1$, $b=2, c=-n-1$ lead respectively to

$$
f(n+1) \in\left\{k(n+1)^{2}, k(n-1)^{2}\right\} \quad \text { and } \quad f(n+1) \in\left\{k(n+1)^{2}, k(n-3)^{2}\right\}
$$

Since $k(n-1)^{2} \neq k(n-3)^{2}$ for $n \neq 2$, the only possibility is $f(n+1)=k(n+1)^{2}$. This completes the induction, so $f(x)=k x^{2}$ for all $x \geq 0$. The same expression is valid for negative values of $x$ since $f$ is even. To verify that $f(x)=k x^{2}$ is actually a solution, we need to check the identity $a^{4}+b^{4}+(a+b)^{4}=2 a^{2} b^{2}+2 a^{2}(a+b)^{2}+2 b^{2}(a+b)^{2}$, which follows directly by expanding both sides.

Therefore the only possible solutions of the functional equation are the constant function $f_{1}(x)=0$ and the following functions:

$$
f_{2}(x)=k x^{2} \quad f_{3}(x)=\left\{\begin{array}{cc}
0 & x \text { even } \\
k & x \text { odd }
\end{array} \quad f_{4}(x)=\left\{\begin{array}{ccc}
0 & x \equiv 0 & (\bmod 4) \\
k & x \equiv 1 & (\bmod 2) \\
4 k & x \equiv 2 & (\bmod 4)
\end{array}\right.\right.
$$

for any non-zero integer $k$. The verification that they are indeed solutions was done for the first two. For $f_{3}$ note that if $a+b+c=0$ then either $a, b, c$ are all even, in which case $f(a)=f(b)=f(c)=0$, or one of them is even and the other two are odd, so both sides of the equation equal $2 k^{2}$. For $f_{4}$ we use similar parity considerations and the symmetry of the equation, which reduces the verification to the triples $(0, k, k),(4 k, k, k),(0,0,0),(0,4 k, 4 k)$. They all satisfy the equation.

Comment. We used several times the same fact: For any $a, b \in \mathbb{Z}$ the functional equation is a quadratic equation in $f(a+b)$ whose coefficients depend on $f(a)$ and $f(b)$ :

$$
f(a+b)^{2}-2(f(a)+f(b)) f(a+b)+(f(a)-f(b))^{2}=0 .
$$

Its discriminant is $16 f(a) f(b)$. Since this value has to be non-negative for any $a, b \in \mathbb{Z}$, we conclude that either $f$ or $-f$ is always non-negative. Also, if $f$ is a solution of the functional equation, then $-f$ is also a solution. Therefore we can assume $f(x) \geq 0$ for all $x \in \mathbb{Z}$. Now, the two solutions of the quadratic equation are

$$
f(a+b) \in\left\{(\sqrt{f(a)}+\sqrt{f(b)})^{2},(\sqrt{f(a)}-\sqrt{f(b)})^{2}\right\} \quad \text { for all } a, b \in \mathbb{Z}
$$

The computation of $f(3)$ from $f(1), f(2)$ and $f(4)$ that we did above follows immediately by setting $(a, b)=(1,2)$ and $(a, b)=(1,-4)$. The inductive step, where $f(n+1)$ is derived from $f(n), f(n-1)$, $f(2)$ and $f(1)$, follows immediately using $(a, b)=(n, 1)$ and $(a, b)=(n-1,2)$.

A2. Let $\mathbb{Z}$ and $\mathbb{Q}$ be the sets of integers and rationals respectively.
a) Does there exist a partition of $\mathbb{Z}$ into three non-empty subsets $A, B, C$ such that the sets $A+B, B+C, C+A$ are disjoint?
b) Does there exist a partition of $\mathbb{Q}$ into three non-empty subsets $A, B, C$ such that the sets $A+B, B+C, C+A$ are disjoint?

Here $X+Y$ denotes the set $\{x+y \mid x \in X, y \in Y\}$, for $X, Y \subseteq \mathbb{Z}$ and $X, Y \subseteq \mathbb{Q}$.
Solution 1. a) The residue classes modulo 3 yield such a partition:

$$
A=\{3 k \mid k \in \mathbb{Z}\}, \quad B=\{3 k+1 \mid k \in \mathbb{Z}\}, \quad C=\{3 k+2 \mid k \in \mathbb{Z}\} .
$$

b) The answer is no. Suppose that $\mathbb{Q}$ can be partitioned into non-empty subsets $A, B, C$ as stated. Note that for all $a \in A, b \in B, c \in C$ one has

$$
\begin{equation*}
a+b-c \in C, \quad b+c-a \in A, \quad c+a-b \in B . \tag{1}
\end{equation*}
$$

Indeed $a+b-c \notin A$ as $(A+B) \cap(A+C)=\emptyset$, and similarly $a+b-c \notin B$, hence $a+b-c \in C$. The other two relations follow by symmetry. Hence $A+B \subset C+C, B+C \subset A+A, C+A \subset B+B$.

The opposite inclusions also hold. Let $a, a^{\prime} \in A$ and $b \in B, c \in C$ be arbitrary. By (1) $a^{\prime}+c-b \in B$, and since $a \in A, c \in C$, we use (1) again to obtain

$$
a+a^{\prime}-b=a+\left(a^{\prime}+c-b\right)-c \in C .
$$

So $A+A \subset B+C$ and likewise $B+B \subset C+A, C+C \subset A+B$. In summary

$$
B+C=A+A, \quad C+A=B+B, \quad A+B=C+C .
$$

Furthermore suppose that $0 \in A$ without loss of generality. Then $B=\{0\}+B \subset A+B$ and $C=\{0\}+C \subset A+C$. So, since $B+C$ is disjoint with $A+B$ and $A+C$, it is also disjoint with $B$ and $C$. Hence $B+C$ is contained in $\mathbb{Z} \backslash(B \cup C)=A$. Because $B+C=A+A$, we obtain $A+A \subset A$. On the other hand $A=\{0\}+A \subset A+A$, implying $A=A+A=B+C$.

Therefore $A+B+C=A+A+A=A$, and now $B+B=C+A$ and $C+C=A+B$ yield $B+B+B=A+B+C=A, C+C+C=A+B+C=A$. In particular if $r \in \mathbb{Q}=A \cup B \cup C$ is arbitrary then $3 r \in A$.

However such a conclusion is impossible. Take any $b \in B(B \neq \emptyset)$ and let $r=b / 3 \in \mathbb{Q}$. Then $b=3 r \in A$ which is a contradiction.

Solution 2. We prove that the example for $\mathbb{Z}$ from the first solution is unique, and then use this fact to solve part b).

Let $\mathbb{Z}=A \cup B \cup C$ be a partition of $\mathbb{Z}$ with $A, B, C \neq \emptyset$ and $A+B, B+C, C+A$ disjoint. We need the relations (1) which clearly hold for $\mathbb{Z}$. Fix two consecutive integers from different sets, say $b \in B$ and $c=b+1 \in C$. For every $a \in A$ we have, in view of (1), $a-1=a+b-c \in C$ and $a+1=a+c-b \in B$. So every $a \in A$ is preceded by a number from $C$ and followed by a number from $B$.

In particular there are pairs of the form $c, c+1$ with $c \in C, c+1 \in A$. For such a pair and any $b \in B$ analogous reasoning shows that each $b \in B$ is preceded by a number from $A$ and followed by a number from $C$. There are also pairs $b, b-1$ with $b \in B, b-1 \in A$. We use them in a similar way to prove that each $c \in C$ is preceded by a number from $B$ and followed by a number from $A$.

By putting the observations together we infer that $A, B, C$ are the three congruence classes modulo 3. Observe that all multiples of 3 are in the set of the partition that contains 0 .

Now we turn to part b). Suppose that there is a partition of $\mathbb{Q}$ with the given properties. Choose three rationals $r_{i}=p_{i} / q_{i}$ from the three sets $A, B, C, i=1,2,3$, and set $N=3 q_{1} q_{2} q_{3}$.

Let $S \subset \mathbb{Q}$ be the set of fractions with denominators $N$ (irreducible or not). It is obtained through multiplication of every integer by the constant $1 / N$, hence closed under sums and differences. Moreover, if we identify each $k \in \mathbb{Z}$ with $k / N \in S$ then $S$ is essentially the set $\mathbb{Z}$ with respect to addition. The numbers $r_{i}$ belong to $S$ because

$$
r_{1}=\frac{3 p_{1} q_{2} q_{3}}{N}, \quad r_{2}=\frac{3 p_{2} q_{3} q_{1}}{N}, \quad r_{3}=\frac{3 p_{3} q_{1} q_{2}}{N}
$$

The partition $\mathbb{Q}=A \cup B \cup C$ of $\mathbb{Q}$ induces a partition $S=A^{\prime} \cup B^{\prime} \cup C^{\prime}$ of $S$, with $A^{\prime}=A \cap S$, $B^{\prime}=B \cap S, C^{\prime}=C \cap S$. Clearly $A^{\prime}+B^{\prime}, B^{\prime}+C^{\prime}, C^{\prime}+A^{\prime}$ are disjoint, so this partition has the properties we consider.

By the uniqueness of the example for $\mathbb{Z}$ the sets $A^{\prime}, B^{\prime}, C^{\prime}$ are the congruence classes modulo 3 , multiplied by $1 / N$. Also all multiples of $3 / N$ are in the same set, $A^{\prime}, B^{\prime}$ or $C^{\prime}$. This holds for $r_{1}, r_{2}, r_{3}$ in particular as they are all multiples of $3 / N$. However $r_{1}, r_{2}, r_{3}$ are in different sets $A^{\prime}, B^{\prime}, C^{\prime}$ since they were chosen from different sets $A, B, C$. The contradiction ends the proof.

Comment. The uniqueness of the example for $\mathbb{Z}$ can also be deduced from the argument in the first solution.

A3. Let $a_{2}, \ldots, a_{n}$ be $n-1$ positive real numbers, where $n \geq 3$, such that $a_{2} a_{3} \cdots a_{n}=1$. Prove that

$$
\left(1+a_{2}\right)^{2}\left(1+a_{3}\right)^{3} \cdots\left(1+a_{n}\right)^{n}>n^{n} .
$$

Solution. The substitution $a_{2}=\frac{x_{2}}{x_{1}}, a_{3}=\frac{x_{3}}{x_{2}}, \ldots, a_{n}=\frac{x_{1}}{x_{n-1}}$ transforms the original problem into the inequality

$$
\begin{equation*}
\left(x_{1}+x_{2}\right)^{2}\left(x_{2}+x_{3}\right)^{3} \cdots\left(x_{n-1}+x_{1}\right)^{n}>n^{n} x_{1}^{2} x_{2}^{3} \cdots x_{n-1}^{n} \tag{*}
\end{equation*}
$$

for all $x_{1}, \ldots, x_{n-1}>0$. To prove this, we use the AM-GM inequality for each factor of the left-hand side as follows:

$$
\begin{array}{rlll}
\left(x_{1}+x_{2}\right)^{2} & & & \geq 2^{2} x_{1} x_{2} \\
\left(x_{2}+x_{3}\right)^{3} & = & \left(2\left(\frac{x_{2}}{2}\right)+x_{3}\right)^{3} & \geq 3^{3}\left(\frac{x_{2}}{2}\right)^{2} x_{3} \\
\left(x_{3}+x_{4}\right)^{4} & = & \left(3\left(\frac{x_{3}}{3}\right)+x_{4}\right)^{4} & \geq 4^{4}\left(\frac{x_{3}}{3}\right)^{3} x_{4} \\
& \vdots & \vdots & \vdots \\
\left(x_{n-1}+x_{1}\right)^{n} & = & \left((n-1)\left(\frac{x_{n-1}}{n-1}\right)+x_{1}\right)^{n} & \geq n^{n}\left(\frac{x_{n-1}}{n-1}\right)^{n-1} x_{1} .
\end{array}
$$

Multiplying these inequalities together gives ( ${ }^{*}$ ), with inequality sign $\geq$ instead of $>$. However for the equality to occur it is necessary that $x_{1}=x_{2}, x_{2}=2 x_{3}, \ldots, x_{n-1}=(n-1) x_{1}$, implying $x_{1}=(n-1)!x_{1}$. This is impossible since $x_{1}>0$ and $n \geq 3$. Therefore the inequality is strict.

Comment. One can avoid the substitution $a_{i}=x_{i} / x_{i-1}$. Apply the weighted AM-GM inequality to each factor $\left(1+a_{k}\right)^{k}$, with the same weights like above, to obtain

$$
\left(1+a_{k}\right)^{k}=\left((k-1) \frac{1}{k-1}+a_{k}\right)^{k} \geq \frac{k^{k}}{(k-1)^{k-1}} a_{k} .
$$

Multiplying all these inequalities together gives

$$
\left(1+a_{2}\right)^{2}\left(1+a_{3}\right)^{3} \cdots\left(1+a_{n}\right)^{n} \geq n^{n} a_{2} a_{3} \cdots a_{n}=n^{n} .
$$

The same argument as in the proof above shows that the equality cannot be attained.

A4. Let $f$ and $g$ be two nonzero polynomials with integer coefficients and $\operatorname{deg} f>\operatorname{deg} g$. Suppose that for infinitely many primes $p$ the polynomial $p f+g$ has a rational root. Prove that $f$ has a rational root.

Solution 1. Since $\operatorname{deg} f>\operatorname{deg} g$, we have $|g(x) / f(x)|<1$ for sufficiently large $x$; more precisely, there is a real number $R$ such that $|g(x) / f(x)|<1$ for all $x$ with $|x|>R$. Then for all such $x$ and all primes $p$ we have

$$
|p f(x)+g(x)| \geq|f(x)|\left(p-\frac{|g(x)|}{|f(x)|}\right)>0 .
$$

Hence all real roots of the polynomials $p f+g$ lie in the interval $[-R, R]$.
Let $f(x)=a_{n} x^{n}+a_{n-1} x^{n-1}+\cdots+a_{0}$ and $g(x)=b_{m} x^{m}+b_{m-1} x^{m-1}+\cdots+b_{0}$ where $n>m, a_{n} \neq 0$ and $b_{m} \neq 0$. Upon replacing $f(x)$ and $g(x)$ by $a_{n}^{n-1} f\left(x / a_{n}\right)$ and $a_{n}^{n-1} g\left(x / a_{n}\right)$ respectively, we reduce the problem to the case $a_{n}=1$. In other words one can assume that $f$ is monic. Then the leading coefficient of $p f+g$ is $p$, and if $r=u / v$ is a rational root of $p f+g$ with $(u, v)=1$ and $v>0$, then either $v=1$ or $v=p$.

First consider the case when $v=1$ infinitely many times. If $v=1$ then $|u| \leq R$, so there are only finitely many possibilities for the integer $u$. Therefore there exist distinct primes $p$ and $q$ for which we have the same value of $u$. Then the polynomials $p f+g$ and $q f+g$ share this root, implying $f(u)=g(u)=0$. So in this case $f$ and $g$ have an integer root in common.

Now suppose that $v=p$ infinitely many times. By comparing the exponent of $p$ in the denominators of $p f(u / p)$ and $g(u / p)$ we get $m=n-1$ and $p f(u / p)+g(u / p)=0$ reduces to an equation of the form

$$
\left(u^{n}+a_{n-1} p u^{n-1}+\ldots+a_{0} p^{n}\right)+\left(b_{n-1} u^{n-1}+b_{n-2} p u^{n-2}+\ldots+b_{0} p^{n-1}\right)=0 .
$$

The equation above implies that $u^{n}+b_{n-1} u^{n-1}$ is divisible by $p$ and hence, since $(u, p)=1$, we have $u+b_{n-1}=p k$ with some integer $k$. On the other hand all roots of $p f+g$ lie in the interval $[-R, R]$, so that

$$
\begin{gathered}
\frac{\left|p k-b_{n-1}\right|}{p}=\frac{|u|}{p}<R, \\
|k|<R+\frac{\left|b_{n-1}\right|}{p}<R+\left|b_{n-1}\right| .
\end{gathered}
$$

Therefore the integer $k$ can attain only finitely many values. Hence there exists an integer $k$ such that the number $\frac{p k-b_{n-1}}{p}=k-\frac{b_{n-1}}{p}$ is a root of $p f+g$ for infinitely many primes $p$. For these primes we have

$$
f\left(k-b_{n-1} \frac{1}{p}\right)+\frac{1}{p} g\left(k-b_{n-1} \frac{1}{p}\right)=0 .
$$

So the equation

$$
\begin{equation*}
f\left(k-b_{n-1} x\right)+x g\left(k-b_{n-1} x\right)=0 \tag{1}
\end{equation*}
$$

has infinitely many solutions of the form $x=1 / p$. Since the left-hand side is a polynomial, this implies that (1) is a polynomial identity, so it holds for all real $x$. In particular, by substituting $x=0$ in (1) we get $f(k)=0$. Thus the integer $k$ is a root of $f$.

In summary the monic polynomial $f$ obtained after the initial reduction always has an integer root. Therefore the original polynomial $f$ has a rational root.

Solution 2. Analogously to the first solution, there exists a real number $R$ such that the complex roots of all polynomials of the form $p f+g$ lie in the disk $|z| \leq R$.

For each prime $p$ such that $p f+g$ has a rational root, by Gauss' lemma $p f+g$ is the product of two integer polynomials, one with degree 1 and the other with degree $\operatorname{deg} f-1$. Since $p$ is a prime, the leading coefficient of one of these factors divides the leading coefficient of $f$. Denote that factor by $h_{p}$.

By narrowing the set of the primes used we can assume that all polynomials $h_{p}$ have the same degree and the same leading coefficient. Their complex roots lie in the disk $|z| \leq R$, hence Vieta's formulae imply that all coefficients of all polynomials $h_{p}$ form a bounded set. Since these coefficients are integers, there are only finitely many possible polynomials $h_{p}$. Hence there is a polynomial $h$ such that $h_{p}=h$ for infinitely many primes $p$.

Finally, if $p$ and $q$ are distinct primes with $h_{p}=h_{q}=h$ then $h$ divides $(p-q) f$. Since $\operatorname{deg} h=1$ or $\operatorname{deg} h=\operatorname{deg} f-1$, in both cases $f$ has a rational root.

Comment. Clearly the polynomial $h$ is a common factor of $f$ and $g$. If $\operatorname{deg} h=1$ then $f$ and $g$ share a rational root. Otherwise $\operatorname{deg} h=\operatorname{deg} f-1$ forces $\operatorname{deg} g=\operatorname{deg} f-1$ and $g$ divides $f$ over the rationals.

Solution 3. Like in the first solution, there is a real number $R$ such that the real roots of all polynomials of the form $p f+g$ lie in the interval $[-R, R]$.

Let $p_{1}<p_{2}<\cdots$ be an infinite sequence of primes so that for every index $k$ the polynomial $p_{k} f+g$ has a rational root $r_{k}$. The sequence $r_{1}, r_{2}, \ldots$ is bounded, so it has a convergent subsequence $r_{k_{1}}, r_{k_{2}}, \ldots$. Now replace the sequences $\left(p_{1}, p_{2}, \ldots\right)$ and $\left(r_{1}, r_{2}, \ldots\right)$ by ( $p_{k_{1}}, p_{k_{2}}, \ldots$ ) and $\left(r_{k_{1}}, r_{k_{2}}, \ldots\right)$; after this we can assume that the sequence $r_{1}, r_{2}, \ldots$ is convergent. Let $\alpha=\lim _{k \rightarrow \infty} r_{k}$. We show that $\alpha$ is a rational root of $f$.

Over the interval $[-R, R]$, the polynomial $g$ is bounded, $|g(x)| \leq M$ with some fixed $M$. Therefore

$$
\left|f\left(r_{k}\right)\right|=\left|f\left(r_{k}\right)-\frac{p_{k} f\left(r_{k}\right)+g\left(r_{k}\right)}{p_{k}}\right|=\frac{\left|g\left(r_{k}\right)\right|}{p_{k}} \leq \frac{M}{p_{k}} \rightarrow 0
$$

and

$$
f(\alpha)=f\left(\lim _{k \rightarrow \infty} r_{k}\right)=\lim _{k \rightarrow \infty} f\left(r_{k}\right)=0
$$

So $\alpha$ is a root of $f$ indeed.
Now let $u_{k}, v_{k}$ be relative prime integers for which $r_{k}=\frac{u_{k}}{v_{k}}$. Let $a$ be the leading coefficient of $f$, let $b=f(0)$ and $c=g(0)$ be the constant terms of $f$ and $g$, respectively. The leading coefficient of the polynomial $p_{k} f+g$ is $p_{k} a$, its constant term is $p_{k} b+c$. So $v_{k}$ divides $p_{k} a$ and $u_{k}$ divides $p_{k} b+c$. Let $p_{k} b+c=u_{k} e_{k}$ (if $p_{k} b+c=u_{k}=0$ then let $e_{k}=1$ ).

We prove that $\alpha$ is rational by using the following fact. Let $\left(p_{n}\right)$ and $\left(q_{n}\right)$ be sequences of integers such that the sequence $\left(p_{n} / q_{n}\right)$ converges. If $\left(p_{n}\right)$ or $\left(q_{n}\right)$ is bounded then $\lim \left(p_{n} / q_{n}\right)$ is rational.

Case 1: There is an infinite subsequence $\left(k_{n}\right)$ of indices such that $v_{k_{n}}$ divides $a$. Then $\left(v_{k_{n}}\right)$ is bounded, so $\alpha=\lim _{n \rightarrow \infty}\left(u_{k_{n}} / v_{k_{n}}\right)$ is rational.

Case 2: There is an infinite subsequence $\left(k_{n}\right)$ of indices such that $v_{k_{n}}$ does not divide $a$. For such indices we have $v_{k_{n}}=p_{k_{n}} d_{k_{n}}$ where $d_{k_{n}}$ is a divisor of $a$. Then

$$
\alpha=\lim _{n \rightarrow \infty} \frac{u_{k_{n}}}{v_{k_{n}}}=\lim _{n \rightarrow \infty} \frac{p_{k_{n}} b+c}{p_{k_{n}} d_{k_{n}} e_{k_{n}}}=\lim _{n \rightarrow \infty} \frac{b}{d_{k_{n}} e_{k_{n}}}+\lim _{n \rightarrow \infty} \frac{c}{p_{k_{n}} d_{k_{n}} e_{k_{n}}}=\lim _{n \rightarrow \infty} \frac{b}{d_{k_{n}} e_{k_{n}}} .
$$

Because the numerator $b$ in the last limit is bounded, $\alpha$ is rational.

A5. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ that satisfy the conditions

$$
f(1+x y)-f(x+y)=f(x) f(y) \quad \text { for all } x, y \in \mathbb{R}
$$

and $f(-1) \neq 0$.
Solution. The only solution is the function $f(x)=x-1, x \in \mathbb{R}$.
We set $g(x)=f(x)+1$ and show that $g(x)=x$ for all real $x$. The conditions take the form

$$
\begin{equation*}
g(1+x y)-g(x+y)=(g(x)-1)(g(y)-1) \quad \text { for all } x, y \in \mathbb{R} \text { and } g(-1) \neq 1 \tag{1}
\end{equation*}
$$

Denote $C=g(-1)-1 \neq 0$. Setting $y=-1$ in (1) gives

$$
\begin{equation*}
g(1-x)-g(x-1)=C(g(x)-1) . \tag{2}
\end{equation*}
$$

Set $x=1$ in (2) to obtain $C(g(1)-1)=0$. Hence $g(1)=1$ as $C \neq 0$. Now plugging in $x=0$ and $x=2$ yields $g(0)=0$ and $g(2)=2$ respectively.

We pass on to the key observations

$$
\begin{array}{ll}
g(x)+g(2-x)=2 & \text { for all } x \in \mathbb{R}, \\
g(x+2)-g(x)=2 & \text { for all } x \in \mathbb{R} . \tag{4}
\end{array}
$$

Replace $x$ by $1-x$ in (2), then change $x$ to $-x$ in the resulting equation. We obtain the relations $g(x)-g(-x)=C(g(1-x)-1), g(-x)-g(x)=C(g(1+x)-1)$. Then adding them up leads to $C(g(1-x)+g(1+x)-2)=0$. Thus $C \neq 0$ implies (3).

Let $u, v$ be such that $u+v=1$. Apply (1) to the pairs $(u, v)$ and $(2-u, 2-v)$ :

$$
g(1+u v)-g(1)=(g(u)-1)(g(v)-1), \quad g(3+u v)-g(3)=(g(2-u)-1)(g(2-v)-1) .
$$

Observe that the last two equations have equal right-hand sides by (3). Hence $u+v=1$ implies

$$
g(u v+3)-g(u v+1)=g(3)-g(1) .
$$

Each $x \leq 5 / 4$ is expressible in the form $x=u v+1$ with $u+v=1$ (the quadratic function $t^{2}-t+(x-1)$ has real roots for $\left.x \leq 5 / 4\right)$. Hence $g(x+2)-g(x)=g(3)-g(1)$ whenever $x \leq 5 / 4$. Because $g(x)=x$ holds for $x=0,1,2$, setting $x=0$ yields $g(3)=3$. This proves (4) for $x \leq 5 / 4$. If $x>5 / 4$ then $-x<5 / 4$ and so $g(2-x)-g(-x)=2$ by the above. On the other hand (3) gives $g(x)=2-g(2-x), g(x+2)=2-g(-x)$, so that $g(x+2)-g(x)=g(2-x)-g(-x)=2$. Thus (4) is true for all $x \in \mathbb{R}$.

Now replace $x$ by $-x$ in (3) to obtain $g(-x)+g(2+x)=2$. In view of (4) this leads to $g(x)+g(-x)=0$, i. e. $g(-x)=-g(x)$ for all $x$. Taking this into account, we apply (1) to the pairs $(-x, y)$ and $(x,-y)$ :
$g(1-x y)-g(-x+y)=(g(x)+1)(1-g(y)), \quad g(1-x y)-g(x-y)=(1-g(x))(g(y)+1)$.
Adding up yields $g(1-x y)=1-g(x) g(y)$. Then $g(1+x y)=1+g(x) g(y)$ by (3). Now the original equation (1) takes the form $g(x+y)=g(x)+g(y)$. Hence $g$ is additive.

By additvity $g(1+x y)=g(1)+g(x y)=1+g(x y)$; since $g(1+x y)=1+g(x) g(y)$ was shown above, we also have $g(x y)=g(x) g(y)$ ( $g$ is multiplicative). In particular $y=x$ gives $g\left(x^{2}\right)=g(x)^{2} \geq 0$ for all $x$, meaning that $g(x) \geq 0$ for $x \geq 0$. Since $g$ is additive and bounded from below on $[0,+\infty)$, it is linear; more exactly $g(x)=g(1) x=x$ for all $x \in \mathbb{R}$.

In summary $f(x)=x-1, x \in \mathbb{R}$. It is straightforward that this function satisfies the requirements.

Comment. There are functions that satisfy the given equation but vanish at -1 , for instance the constant function 0 and $f(x)=x^{2}-1, x \in \mathbb{R}$.

A6. Let $f: \mathbb{N} \rightarrow \mathbb{N}$ be a function, and let $f^{m}$ be $f$ applied $m$ times. Suppose that for every $n \in \mathbb{N}$ there exists a $k \in \mathbb{N}$ such that $f^{2 k}(n)=n+k$, and let $k_{n}$ be the smallest such $k$. Prove that the sequence $k_{1}, k_{2}, \ldots$ is unbounded.

Solution. We restrict attention to the set

$$
S=\left\{1, f(1), f^{2}(1), \ldots\right\}
$$

Observe that $S$ is unbounded because for every number $n$ in $S$ there exists a $k>0$ such that $f^{2 k}(n)=n+k$ is in $S$. Clearly $f$ maps $S$ into itself; moreover $f$ is injective on $S$. Indeed if $f^{i}(1)=f^{j}(1)$ with $i \neq j$ then the values $f^{m}(1)$ start repeating periodically from some point on, and $S$ would be finite.

Define $g: S \rightarrow S$ by $g(n)=f^{2 k_{n}}(n)=n+k_{n}$. We prove that $g$ is injective too. Suppose that $g(a)=g(b)$ with $a<b$. Then $a+k_{a}=f^{2 k_{a}}(a)=f^{2 k_{b}}(b)=b+k_{b}$ implies $k_{a}>k_{b}$. So, since $f$ is injective on $S$, we obtain

$$
f^{2\left(k_{a}-k_{b}\right)}(a)=b=a+\left(k_{a}-k_{b}\right) .
$$

However this contradicts the minimality of $k_{a}$ as $0<k_{a}-k_{b}<k_{a}$.
Let $T$ be the set of elements of $S$ that are not of the form $g(n)$ with $n \in S$. Note that $1 \in T$ by $g(n)>n$ for $n \in S$, so $T$ is non-empty. For each $t \in T$ denote $C_{t}=\left\{t, g(t), g^{2}(t), \ldots\right\}$; call $C_{t}$ the chain starting at $t$. Observe that distinct chains are disjoint because $g$ is injective. Each $n \in S \backslash T$ has the form $n=g\left(n^{\prime}\right)$ with $n^{\prime}<n, n^{\prime} \in S$. Repeated applications of the same observation show that $n \in C_{t}$ for some $t \in T$, i. e. $S$ is the disjoint union of the chains $C_{t}$.

If $f^{n}(1)$ is in the chain $C_{t}$ starting at $t=f^{n_{t}}(1)$ then $n=n_{t}+2 a_{1}+\cdots+2 a_{j}$ with

$$
f^{n}(1)=g^{j}\left(f^{n_{t}}(1)\right)=f^{2 a_{j}}\left(f^{2 a_{j-1}}\left(\cdots f^{2 a_{1}}\left(f^{n_{t}}(1)\right)\right)\right)=f^{n_{t}}(1)+a_{1}+\cdots+a_{j} .
$$

Hence

$$
\begin{equation*}
f^{n}(1)=f^{n_{t}}(1)+\frac{n-n_{t}}{2}=t+\frac{n-n_{t}}{2} . \tag{1}
\end{equation*}
$$

Now we show that $T$ is infinite. We argue by contradiction. Suppose that there are only finitely many chains $C_{t_{1}}, \ldots, C_{t_{r}}$, starting at $t_{1}<\cdots<t_{r}$. Fix $N$. If $f^{n}(1)$ with $1 \leq n \leq N$ is in $C_{t}$ then $f^{n}(1)=t+\frac{n-n_{t}}{2} \leq t_{r}+\frac{N}{2}$ by (1). But then the $N+1$ distinct natural numbers $1, f(1), \ldots, f^{N}(1)$ are all less than $t_{r}+\frac{N}{2}$ and hence $N+1 \leq t_{r}+\frac{N}{2}$. This is a contradiction if $N$ is sufficiently large, and hence $T$ is infinite.

To complete the argument, choose any $k$ in $\mathbb{N}$ and consider the $k+1$ chains starting at the first $k+1$ numbers in $T$. Let $t$ be the greatest one among these numbers. Then each of the chains in question contains a number not exceeding $t$, and at least one of them does not contain any number among $t+1, \ldots, t+k$. So there is a number $n$ in this chain such that $g(n)-n>k$, i. e. $k_{n}>k$. In conclusion $k_{1}, k_{2}, \ldots$ is unbounded.

A7. We say that a function $f: \mathbb{R}^{k} \rightarrow \mathbb{R}$ is a metapolynomial if, for some positive integers $m$ and $n$, it can be represented in the form

$$
f\left(x_{1}, \ldots, x_{k}\right)=\max _{i=1, \ldots, m} \min _{j=1, \ldots, n} P_{i, j}\left(x_{1}, \ldots, x_{k}\right)
$$

where $P_{i, j}$ are multivariate polynomials. Prove that the product of two metapolynomials is also a metapolynomial.

Solution. We use the notation $f(x)=f\left(x_{1}, \ldots, x_{k}\right)$ for $x=\left(x_{1}, \ldots, x_{k}\right)$ and $[m]=\{1,2, \ldots, m\}$. Observe that if a metapolynomial $f(x)$ admits a representation like the one in the statement for certain positive integers $m$ and $n$, then they can be replaced by any $m^{\prime} \geq m$ and $n^{\prime} \geq n$. For instance, if we want to replace $m$ by $m+1$ then it is enough to define $P_{m+1, j}(x)=P_{m, j}(x)$ and note that repeating elements of a set do not change its maximum nor its minimum. So one can assume that any two metapolynomials are defined with the same $m$ and $n$. We reserve letters $P$ and $Q$ for polynomials, so every function called $P, P_{i, j}, Q, Q_{i, j}, \ldots$ is a polynomial function.

We start with a lemma that is useful to change expressions of the form $\min \max f_{i, j}$ to ones of the form max min $g_{i, j}$.
Lemma. Let $\left\{a_{i, j}\right\}$ be real numbers, for all $i \in[m]$ and $j \in[n]$. Then

$$
\min _{i \in[m]} \max _{j \in[n]} a_{i, j}=\max _{j_{1}, \ldots, j_{m} \in[n]} \min _{i \in[m]} a_{i, j_{i}}
$$

where the max in the right-hand side is over all vectors $\left(j_{1}, \ldots, j_{m}\right)$ with $j_{1}, \ldots, j_{m} \in[n]$.
Proof. We can assume for all $i$ that $a_{i, n}=\max \left\{a_{i, 1}, \ldots, a_{i, n}\right\}$ and $a_{m, n}=\min \left\{a_{1, n}, \ldots, a_{m, n}\right\}$. The left-hand side is $=a_{m, n}$ and hence we need to prove the same for the right-hand side. If $\left(j_{1}, j_{2}, \ldots, j_{m}\right)=(n, n, \ldots, n)$ then $\min \left\{a_{1, j_{1}}, \ldots, a_{m, j_{m}}\right\}=\min \left\{a_{1, n}, \ldots, a_{m, n}\right\}=a_{m, n}$ which implies that the right-hand side is $\geq a_{m, n}$. It remains to prove the opposite inequality and this is equivalent to $\min \left\{a_{1, j_{1}}, \ldots, a_{m, j_{m}}\right\} \leq a_{m, n}$ for all possible $\left(j_{1}, j_{2}, \ldots, j_{m}\right)$. This is true because $\min \left\{a_{1, j_{1}}, \ldots, a_{m, j_{m}}\right\} \leq a_{m, j_{m}} \leq a_{m, n}$.

We need to show that the family $\mathcal{M}$ of metapolynomials is closed under multiplication, but it turns out easier to prove more: that it is also closed under addition, maxima and minima.

First we prove the assertions about the maxima and the minima. If $f_{1}, \ldots, f_{r}$ are metapolynomials, assume them defined with the same $m$ and $n$. Then

$$
f=\max \left\{f_{1}, \ldots, f_{r}\right\}=\max \left\{\max _{i \in[m]} \min _{j \in[n]} P_{i, j}^{1}, \ldots, \max _{i \in[m]} \min _{j \in[n]} P_{i, j}^{r}\right\}=\max _{s \in[r], i \in[m]} \min _{j \in[n]} P_{i, j}^{s}
$$

It follows that $f=\max \left\{f_{1}, \ldots, f_{r}\right\}$ is a metapolynomial. The same argument works for the minima, but first we have to replace $\min \max$ by $\max \min$, and this is done via the lemma.

Another property we need is that if $f=\max \min P_{i, j}$ is a metapolynomial then so is $-f$. Indeed, $-f=\min \left(-\min P_{i, j}\right)=\min \max P_{i, j}$.

To prove $\mathcal{M}$ is closed under addition let $f=\max \min P_{i, j}$ and $g=\max \min Q_{i, j}$. Then

$$
\begin{gathered}
f(x)+g(x)=\max _{i \in[m]} \min _{j \in[n]} P_{i, j}(x)+\max _{i \in[m]} \min _{j \in[n]} Q_{i, j}(x) \\
=\max _{i_{1}, i_{2} \in[m]}\left(\min _{j \in[n]} P_{i_{1}, j}(x)+\min _{j \in[n]} Q_{i_{2}, j}(x)\right)=\max _{i_{1}, i_{2} \in[m]} \min _{j_{1}, j_{2} \in[n]}\left(P_{i_{1}, j_{1}}(x)+Q_{i_{2}, j_{2}}(x)\right),
\end{gathered}
$$

and hence $f(x)+g(x)$ is a metapolynomial.
We proved that $\mathcal{M}$ is closed under sums, maxima and minima, in particular any function that can be expressed by sums, max, min, polynomials or even metapolynomials is in $\mathcal{M}$.

We would like to proceed with multiplication along the same lines like with addition, but there is an essential difference. In general the product of the maxima of two sets is not equal
to the maximum of the product of the sets. We need to deal with the fact that $a<b$ and $c<d$ do not imply $a c<b d$. However this is true for $a, b, c, d \geq 0$.

In view of this we decompose each function $f(x)$ into its positive part $f^{+}(x)=\max \{f(x), 0\}$ and its negative part $f^{-}(x)=\max \{0,-f(x)\}$. Note that $f=f^{+}-f^{-}$and $f^{+}, f^{-} \in \mathcal{M}$ if $f \in \mathcal{M}$. The whole problem reduces to the claim that if $f$ and $g$ are metapolynomials with $f, g \geq 0$ then $f g$ it is also a metapolynomial.

Assuming this claim, consider arbitrary $f, g \in \mathcal{M}$. We have

$$
f g=\left(f^{+}-f^{-}\right)\left(g^{+}-g^{-}\right)=f^{+} g^{+}-f^{+} g^{-}-f^{-} g^{+}+f^{-} g^{-},
$$

and hence $f g \in \mathcal{M}$. Indeed, $\mathcal{M}$ is closed under addition, also $f^{+} g^{+}, f^{+} g^{-}, f^{-} g^{+}, f^{-} g^{-} \in \mathcal{M}$ because $f^{+}, f^{-}, g^{+}, g^{-} \geq 0$.

It remains to prove the claim. In this case $f, g \geq 0$, and one can try to repeat the argument for the sum. More precisely, let $f=\max \min P_{i j} \geq 0$ and $g=\max \min Q_{i j} \geq 0$. Then

$$
f g=\max \min P_{i, j} \cdot \max \min Q_{i, j}=\max \min P_{i, j}^{+} \cdot \max \min Q_{i, j}^{+}=\max \min P_{i_{1}, j_{1}}^{+} \cdot Q_{i_{2}, j_{2}}^{+} .
$$

Hence it suffices to check that $P^{+} Q^{+} \in \mathcal{M}$ for any pair of polynomials $P$ and $Q$. This reduces to the identity

$$
u^{+} v^{+}=\max \left\{0, \min \{u v, u, v\}, \min \left\{u v, u v^{2}, u^{2} v\right\}, \min \left\{u v, u, u^{2} v\right\}, \min \left\{u v, u v^{2}, v\right\}\right\},
$$

with $u$ replaced by $P(x)$ and $v$ replaced by $Q(x)$. The formula is proved by a case-by-case analysis. If $u \leq 0$ or $v \leq 0$ then both sides equal 0 . In case $u, v \geq 0$, the right-hand side is clearly $\leq u v$. To prove the opposite inequality we use that $u v$ equals

$$
\begin{array}{ll}
\min \{u v, u, v\} & \text { if } 0 \leq u, v \leq 1, \\
\min \left\{u v, u v^{2}, u^{2} v\right\} & \text { if } 1 \leq u, v, \\
\min \left\{u v, u, u^{2} v\right\} & \text { if } 0 \leq v \leq 1 \leq u, \\
\min \left\{u v, u v^{2}, v\right\} & \text { if } 0 \leq u \leq 1 \leq v .
\end{array}
$$

Comment. The case $k=1$ is simpler and can be solved by proving that a function $f: \mathbb{R} \rightarrow \mathbb{R}$ is a metapolynomial if and only if it is a piecewise polinomial (and continuos) function.

It is enough to prove that all such functions are metapolynomials, and this easily reduces to the following case. Given a polynomial $P(x)$ with $P(0)=0$, the function $f$ defined by $f(x)=P(x)$ for $x \geq 0$ and 0 otherwise is a metapolynomial. For this last claim, it suffices to prove that $\left(x^{+}\right)^{n}$ is a metapolynomial, and this follows from the formula $\left(x^{+}\right)^{n}=\max \left\{0, \min \left\{x^{n-1}, x^{n}\right\}, \min \left\{x^{n}, x^{n+1}\right\}\right\}$.

## Combinatorics

C1. Several positive integers are written in a row. Iteratively, Alice chooses two adjacent numbers $x$ and $y$ such that $x>y$ and $x$ is to the left of $y$, and replaces the pair $(x, y)$ by either $(y+1, x)$ or $(x-1, x)$. Prove that she can perform only finitely many such iterations.

Solution 1. Note first that the allowed operation does not change the maximum $M$ of the initial sequence. Let $a_{1}, a_{2}, \ldots, a_{n}$ be the numbers obtained at some point of the process. Consider the sum

$$
S=a_{1}+2 a_{2}+\cdots+n a_{n} .
$$

We claim that $S$ increases by a positive integer amount with every operation. Let the operation replace the pair $\left(a_{i}, a_{i+1}\right)$ by a pair $\left(c, a_{i}\right)$, where $a_{i}>a_{i+1}$ and $c=a_{i+1}+1$ or $c=a_{i}-1$. Then the new and the old value of $S$ differ by $d=\left(i c+(i+1) a_{i}\right)-\left(i a_{i}+(i+1) a_{i+1}\right)=a_{i}-a_{i+1}+i\left(c-a_{i+1}\right)$. The integer $d$ is positive since $a_{i}-a_{i+1} \geq 1$ and $c-a_{i+1} \geq 0$.

On the other hand $S \leq(1+2+\cdots+n) M$ as $a_{i} \leq M$ for all $i=1, \ldots, n$. Since $S$ increases by at least 1 at each step and never exceeds the constant $(1+2+\cdots+n) M$, the process stops after a finite number of iterations.

Solution 2. Like in the first solution note that the operations do not change the maximum $M$ of the initial sequence. Now consider the reverse lexicographical order for $n$-tuples of integers. We say that $\left(x_{1}, \ldots, x_{n}\right)<\left(y_{1}, \ldots, y_{n}\right)$ if $x_{n}<y_{n}$, or if $x_{n}=y_{n}$ and $x_{n-1}<y_{n-1}$, or if $x_{n}=y_{n}$, $x_{n-1}=y_{n-1}$ and $x_{n-2}<y_{n-2}$, etc. Each iteration creates a sequence that is greater than the previous one with respect to this order, and no sequence occurs twice during the process. On the other hand there are finitely many possible sequences because their terms are always positive integers not exceeding $M$. Hence the process cannot continue forever.

Solution 3. Let the current numbers be $a_{1}, a_{2}, \ldots, a_{n}$. Define the score $s_{i}$ of $a_{i}$ as the number of $a_{j}$ 's that are less than $a_{i}$. Call the sequence $s_{1}, s_{2}, \ldots, s_{n}$ the score sequence of $a_{1}, a_{2}, \ldots, a_{n}$.

Let us say that a sequence $x_{1}, \ldots, x_{n}$ dominates a sequence $y_{1}, \ldots, y_{n}$ if the first index $i$ with $x_{i} \neq y_{i}$ is such that $x_{i}<y_{i}$. We show that after each operation the new score sequence dominates the old one. Score sequences do not repeat, and there are finitely many possibilities for them, no more than $(n-1)^{n}$. Hence the process will terminate.

Consider an operation that replaces $(x, y)$ by $(a, x)$, with $a=y+1$ or $a=x-1$. Suppose that $x$ was originally at position $i$. For each $j<i$ the score $s_{j}$ does not increase with the change because $y \leq a$ and $x \leq x$. If $s_{j}$ decreases for some $j<i$ then the new score sequence dominates the old one. Assume that $s_{j}$ stays the same for all $j<i$ and consider $s_{i}$. Since $x>y$ and $y \leq a \leq x$, we see that $s_{i}$ decreases by at least 1 . This concludes the proof.

Comment. All three proofs work if $x$ and $y$ are not necessarily adjacent, and if the pair $(x, y)$ is replaced by any pair ( $a, x$ ), with $a$ an integer satisfying $y \leq a \leq x$. There is nothing special about the "weights" $1,2, \ldots, n$ in the definition of $S=\sum_{i=1}^{n} i a_{i}$ from the first solution. For any sequence $w_{1}<w_{2}<\cdots<w_{n}$ of positive integers, the sum $\sum_{i=1}^{n} w_{i} a_{i}$ increases by at least 1 with each operation.

Consider the same problem, but letting Alice replace the pair $(x, y)$ by $(a, x)$, where $a$ is any positive integer less than $x$. The same conclusion holds in this version, i. e. the process stops eventually. The solution using the reverse lexicographical order works without any change. The first solution would require a special set of weights like $w_{i}=M^{i}$ for $i=1, \ldots, n$.

Comment. The first and the second solutions provide upper bounds for the number of possible operations, respectively of order $M n^{2}$ and $M^{n}$ where $M$ is the maximum of the original sequence. The upper bound $(n-1)^{n}$ in the third solution does not depend on $M$.

C2. Let $n \geq 1$ be an integer. What is the maximum number of disjoint pairs of elements of the set $\{1,2, \ldots, n\}$ such that the sums of the different pairs are different integers not exceeding $n$ ?

Solution. Consider $x$ such pairs in $\{1,2, \ldots, n\}$. The sum $S$ of the $2 x$ numbers in them is at least $1+2+\cdots+2 x$ since the pairs are disjoint. On the other hand $S \leq n+(n-1)+\cdots+(n-x+1)$ because the sums of the pairs are different and do not exceed $n$. This gives the inequality

$$
\frac{2 x(2 x+1)}{2} \leq n x-\frac{x(x-1)}{2},
$$

which leads to $x \leq \frac{2 n-1}{5}$. Hence there are at most $\left\lfloor\frac{2 n-1}{5}\right\rfloor$ pairs with the given properties.
We show a construction with exactly $\left\lfloor\frac{2 n-1}{5}\right\rfloor$ pairs. First consider the case $n=5 k+3$ with $k \geq 0$, where $\left\lfloor\frac{2 n-1}{5}\right\rfloor=2 k+1$. The pairs are displayed in the following table.

| Pairs | $3 k+1$ | $3 k$ | $\cdots$ | $2 k+2$ | $4 k+2$ | $4 k+1$ | $\cdots$ | $3 k+3$ | $3 k+2$ |
| :---: | :---: | :---: | :--- | :---: | :---: | :---: | :---: | :---: | :---: |
|  | 2 | 4 | $\cdots$ | $2 k$ | 1 | 3 | $\cdots$ | $2 k-1$ | $2 k+1$ |
| Sums | $3 k+3$ | $3 k+4$ | $\cdots$ | $4 k+2$ | $4 k+3$ | $4 k+4$ | $\cdots$ | $5 k+2$ | $5 k+3$ |

The $2 k+1$ pairs involve all numbers from 1 to $4 k+2$; their sums are all numbers from $3 k+3$ to $5 k+3$. The same construction works for $n=5 k+4$ and $n=5 k+5$ with $k \geq 0$. In these cases the required number $\left\lfloor\frac{2 n-1}{5}\right\rfloor$ of pairs equals $2 k+1$ again, and the numbers in the table do not exceed $5 k+3$. In the case $n=5 k+2$ with $k \geq 0$ one needs only $2 k$ pairs. They can be obtained by ignoring the last column of the table (thus removing $5 k+3$ ). Finally, $2 k$ pairs are also needed for the case $n=5 k+1$ with $k \geq 0$. Now it suffices to ignore the last column of the table and then subtract 1 from each number in the first row.

Comment. The construction above is not unique. For instance, the following table shows another set of $2 k+1$ pairs for the cases $n=5 k+3, n=5 k+4$, and $n=5 k+5$.

| Pairs | 1 | 2 | $\cdots$ | $k$ | $k+1$ | $k+2$ | $\cdots$ | $2 k+1$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  | $4 k+1$ | $4 k-1$ | $\cdots$ | $2 k+3$ | $4 k+2$ | $4 k$ | $\cdots$ | $2 k+2$ |
| Sums | $4 k+2$ | $4 k+1$ | $\cdots$ | $3 k+3$ | $5 k+3$ | $5 k+2$ | $\cdots$ | $4 k+3$ |

The table for the case $n=5 k+2$ would be the same, with the pair $(k+1,4 k+2)$ removed. For the case $n=5 k+1$ remove the last column and subtract 2 from each number in the second row.

C3. In a $999 \times 999$ square table some cells are white and the remaining ones are red. Let $T$ be the number of triples ( $C_{1}, C_{2}, C_{3}$ ) of cells, the first two in the same row and the last two in the same column, with $C_{1}$ and $C_{3}$ white and $C_{2}$ red. Find the maximum value $T$ can attain.

Solution. We prove that in an $n \times n$ square table there are at most $\frac{4 n^{4}}{27}$ such triples.
Let row $i$ and column $j$ contain $a_{i}$ and $b_{j}$ white cells respectively, and let $R$ be the set of red cells. For every red cell $(i, j)$ there are $a_{i} b_{j}$ admissible triples $\left(C_{1}, C_{2}, C_{3}\right)$ with $C_{2}=(i, j)$, therefore

$$
T=\sum_{(i, j) \in R} a_{i} b_{j} .
$$

We use the inequality $2 a b \leq a^{2}+b^{2}$ to obtain

$$
T \leq \frac{1}{2} \sum_{(i, j) \in R}\left(a_{i}^{2}+b_{j}^{2}\right)=\frac{1}{2} \sum_{i=1}^{n}\left(n-a_{i}\right) a_{i}^{2}+\frac{1}{2} \sum_{j=1}^{n}\left(n-b_{j}\right) b_{j}^{2} .
$$

This is because there are $n-a_{i}$ red cells in row $i$ and $n-b_{j}$ red cells in column $j$. Now we maximize the right-hand side.

By the AM-GM inequality we have

$$
(n-x) x^{2}=\frac{1}{2}(2 n-2 x) \cdot x \cdot x \leq \frac{1}{2}\left(\frac{2 n}{3}\right)^{3}=\frac{4 n^{3}}{27}
$$

with equality if and only if $x=\frac{2 n}{3}$. By putting everything together, we get

$$
T \leq \frac{n}{2} \frac{4 n^{3}}{27}+\frac{n}{2} \frac{4 n^{3}}{27}=\frac{4 n^{4}}{27}
$$

If $n=999$ then any coloring of the square table with $x=\frac{2 n}{3}=666$ white cells in each row and column attains the maximum as all inequalities in the previous argument become equalities. For example color a cell $(i, j)$ white if $i-j \equiv 1,2, \ldots, 666(\bmod 999)$, and red otherwise.

Therefore the maximum value $T$ can attain is $T=\frac{4.999^{4}}{27}$.
Comment. One can obtain a better preliminary estimate with the Cauchy-Schwarz inequality:

$$
T=\sum_{(i, j) \in R} a_{i} b_{j} \leq\left(\sum_{(i, j) \in R} a_{i}^{2}\right)^{\frac{1}{2}} \cdot\left(\sum_{(i, j) \in R} b_{j}^{2}\right)^{\frac{1}{2}}=\left(\sum_{i=1}^{n}\left(n-a_{i}\right) a_{i}^{2}\right)^{\frac{1}{2}} \cdot\left(\sum_{j=1}^{n}\left(n-b_{j}\right) b_{j}^{2}\right)^{\frac{1}{2}}
$$

It can be used to reach the same conclusion.

C4. Players $A$ and $B$ play a game with $N \geq 2012$ coins and 2012 boxes arranged around a circle. Initially $A$ distributes the coins among the boxes so that there is at least 1 coin in each box. Then the two of them make moves in the order $B, A, B, A, \ldots$ by the following rules:

- On every move of his $B$ passes 1 coin from every box to an adjacent box.
- On every move of hers $A$ chooses several coins that were not involved in $B$ 's previous move and are in different boxes. She passes every chosen coin to an adjacent box.

Player $A$ 's goal is to ensure at least 1 coin in each box after every move of hers, regardless of how $B$ plays and how many moves are made. Find the least $N$ that enables her to succeed.

Solution. We argue for a general $n \geq 7$ instead of 2012 and prove that the required minimum $N$ is $2 n-2$. For $n=2012$ this gives $N_{\text {min }}=4022$.
a) If $N=2 n-2$ player $A$ can achieve her goal. Let her start the game with a regular distribution: $n-2$ boxes with 2 coins and 2 boxes with 1 coin. Call the boxes of the two kinds red and white respectively. We claim that on her first move $A$ can achieve a regular distribution again, regardless of $B$ 's first move $M$. She acts according as the following situation $S$ occurs after $M$ or not: The initial distribution contains a red box $R$ with 2 white neighbors, and $R$ receives no coins from them on move $M$.

Suppose that $S$ does not occur. Exactly one of the coins $c_{1}$ and $c_{2}$ in a given red box $X$ is involved in $M$, say $c_{1}$. If $M$ passes $c_{1}$ to the right neighbor of $X$, let $A$ pass $c_{2}$ to its left neighbor, and vice versa. By doing so with all red boxes $A$ performs a legal move $M^{\prime}$. Thus $M$ and $M^{\prime}$ combined move the 2 coins of every red box in opposite directions. Hence after $M$ and $M^{\prime}$ are complete each neighbor of a red box $X$ contains exactly 1 coin that was initially in $X$. So each box with a red neighbor is non-empty after $M^{\prime}$. If initially there is a box $X$ with 2 white neighbors ( $X$ is red and unique) then $X$ receives a coin from at least one of them on move $M$ since $S$ does not occur. Such a coin is not involved in $M^{\prime}$, so $X$ is also non-empty after $M^{\prime}$. Furthermore each box $Y$ has given away its initial content after $M$ and $M^{\prime}$. A red neighbor of $Y$ adds 1 coin to it; a white neighbor adds at most 1 coin because it is not involved in $M^{\prime}$. Hence each box contains 1 or 2 coins after $M^{\prime}$. Because $N=2 n-2$, such a distribution is regular.

Now let $S$ occur after move $M$. Then $A$ leaves untouched the exceptional red box $R$. With all remaining red boxes she proceeds like in the previous case, thus making a legal move $M^{\prime \prime}$. Box $R$ receives no coins from its neighbors on either move, so there is 1 coin in it after $M^{\prime \prime}$. Like above $M$ and $M^{\prime \prime}$ combined pass exactly 1 coin from every red box different from $R$ to each of its neighbors. Every box except $R$ has a red neighbor different from $R$, hence all boxes are non-empty after $M^{\prime \prime}$. Next, each box $Y$ except $R$ loses its initial content after $M$ and $M^{\prime \prime}$. A red neighbor of $Y$ adds at most 1 coin to it; a white neighbor also adds at most 1 coin as it does not participate in $M^{\prime \prime}$. Thus each box has 1 or 2 coins after $M^{\prime \prime}$, and the obtained distribution is regular.

Player $A$ can apply the described strategy indefinitely, so $N=2 n-2$ enables her to succeed.
b) For $N \leq 2 n-3$ player $B$ can achieve an empty box after some move of $A$. Let $\alpha$ be a set of $\ell$ consecutive boxes containing a total of $N(\alpha)$ coins. We call $\alpha$ an $\operatorname{arc}$ if $\ell \leq n-2$ and $N(\alpha) \leq 2 \ell-3$. Note that $\ell \geq 2$ by the last condition. Moreover if both extremes of $\alpha$ are non-empty boxes then $N(\alpha) \geq 2$, so that $N(\alpha) \leq 2 \ell-3$ implies $\ell \geq 3$. Observe also that if an extreme $X$ of $\alpha$ has more than 1 coin then ignoring $X$ yields a shorter arc. It follows that every arc contains an arc whose extremes have at most 1 coin each.

Given a clockwise labeling $1,2, \ldots, n$ of the boxes, suppose that boxes $1,2, \ldots, \ell$ form an arc $\alpha$, with $\ell \leq n-2$ and $N(\alpha) \leq 2 \ell-3$. Suppose also that all $n \geq 7$ boxes are non-empty. Then $B$ can move so that an arc $\alpha^{\prime}$ with $N\left(\alpha^{\prime}\right)<N(\alpha)$ will appear after any response of $A$.

One may assume exactly 1 coin in boxes 1 and $\ell$ by a previous remark. Let $B$ pass 1 coin in counterclockwise direction from box 1 and box $n$, and in clockwise direction from each remaining box. This leaves $N(\alpha)-2$ coins in the boxes of $\alpha$. In addition, due to $3 \leq \ell \leq n-2$, box $\ell$ has exactly 1 coin $c$, the one received from box $\ell-1$.

Let player $A$ 's next move $M$ pass $k \leq 2$ coins to boxes $1,2, \ldots, \ell$ from the remaining ones. Only boxes 1 and $\ell$ can receive such coins, at most 1 each. If $k<2$ then after move $M$ boxes $1,2, \ldots, \ell$ form an arc $\alpha^{\prime}$ with $N\left(\alpha^{\prime}\right)<N(\alpha)$. If $k=2$ then $M$ adds a coin to box $\ell$. Also $M$ does not move coin $c$ from $\ell$ because $c$ is involved in the previous move of $B$. In summary boxes $1,2, \ldots, \ell$ contain $N(\alpha)$ coins like before, so they form an arc. However there are 2 coins now in the extreme $\ell$ of the arc. Ignore $\ell$ to obtain a shorter arc $\alpha^{\prime}$ with $N\left(\alpha^{\prime}\right)<N(\alpha)$.

Consider any initial distribution without empty boxes. Since $N \leq 2 n-3$, there are at least 3 boxes in it with exactly 1 coin. It follows from $n \geq 7$ that some 2 of them are the extremes of an arc $\alpha$. Hence $B$ can make the move described above, which leads to an arc $\alpha^{\prime}$ with $N\left(\alpha^{\prime}\right)<N(\alpha)$ after $A^{\prime}$ 's response. If all boxes in the new distribution are non-empty he can repeat the same, and so on. Because $N(\alpha)$ cannot decrease indefinitely, an empty box will occur after some move of $A$.

C5. The columns and the rows of a $3 n \times 3 n$ square board are numbered $1,2, \ldots, 3 n$. Every square $(x, y)$ with $1 \leq x, y \leq 3 n$ is colored asparagus, byzantium or citrine according as the modulo 3 remainder of $x+y$ is 0,1 or 2 respectively. One token colored asparagus, byzantium or citrine is placed on each square, so that there are $3 n^{2}$ tokens of each color.

Suppose that one can permute the tokens so that each token is moved to a distance of at most $d$ from its original position, each asparagus token replaces a byzantium token, each byzantium token replaces a citrine token, and each citrine token replaces an asparagus token. Prove that it is possible to permute the tokens so that each token is moved to a distance of at most $d+2$ from its original position, and each square contains a token with the same color as the square.

Solution. Without loss of generality it suffices to prove that the A-tokens can be moved to distinct A-squares in such a way that each A-token is moved to a distance at most $d+2$ from its original place. This means we need a perfect matching between the $3 n^{2} \mathrm{~A}$-squares and the $3 n^{2}$ A-tokens such that the distance in each pair of the matching is at most $d+2$.

To find the matching, we construct a bipartite graph. The A-squares will be the vertices in one class of the graph; the vertices in the other class will be the A-tokens.

Split the board into $3 \times 1$ horizontal triminos; then each trimino contains exactly one Asquare. Take a permutation $\pi$ of the tokens which moves A-tokens to B-tokens, B-tokens to C-tokens, and C-tokens to A-tokens, in each case to a distance at most $d$. For each A-square $S$, and for each A-token $T$, connect $S$ and $T$ by an edge if $T, \pi(T)$ or $\pi^{-1}(T)$ is on the trimino containing $S$. We allow multiple edges; it is even possible that the same square and the same token are connected with three edges. Obviously the lengths of the edges in the graph do not exceed $d+2$. By length of an edge we mean the distance between the A -square and the A -token it connects.

Each A-token $T$ is connected with the three A-squares whose triminos contain $T, \pi(T)$ and $\pi^{-1}(T)$. Therefore in the graph all tokens are of degree 3 . We show that the same is true for the A-squares. Let $S$ be an arbitrary A-square, and let $T_{1}, T_{2}, T_{3}$ be the three tokens on the trimino containing $S$. For $i=1,2,3$, if $T_{i}$ is an A-token, then $S$ is connected with $T_{i}$; if $T_{i}$ is a B-token then $S$ is connected with $\pi^{-1}\left(T_{i}\right)$; finally, if $T_{i}$ is a C-token then $S$ is connected with $\pi\left(T_{i}\right)$. Hence in the graph the A-squares also are of degree 3 .

Since the A-squares are of degree 3 , from every set $\mathcal{S}$ of A-squares exactly $3|\mathcal{S}|$ edges start. These edges end in at least $|\mathcal{S}|$ tokens because the A-tokens also are of degree 3. Hence every set $\mathcal{S}$ of A -squares has at least $|\mathcal{S}|$ neighbors among the A-tokens.

Therefore, by HALL's marriage theorem, the graph contains a perfect matching between the two vertex classes. So there is a perfect matching between the A -squares and A -tokens with edges no longer than $d+2$. It follows that the tokens can be permuted as specified in the problem statement.

Comment 1. In the original problem proposal the board was infinite and there were only two colors. Having $n$ colors for some positive integer $n$ was an option; we chose $n=3$. Moreover, we changed the board to a finite one to avoid dealing with infinite graphs (although Hall's theorem works in the infinite case as well).

With only two colors Hall's theorem is not needed. In this case we split the board into $2 \times 1$ dominos, and in the resulting graph all vertices are of degree 2 . The graph consists of disjoint cycles with even length and infinite paths, so the existence of the matching is trivial.

Having more than three colors would make the problem statement more complicated, because we need a matching between every two color classes of tokens. However, this would not mean a significant increase in difficulty.

Comment 2. According to Wikipedia, the color asparagus (hexadecimal code \#87A96B) is a tone of green that is named after the vegetable. Crayola created this color in 1993 as one of the 16 to be named in the Name The Color Contest. Byzantium (\#702963) is a dark tone of purple. Its first recorded use as a color name in English was in 1926. Citrine (\#E4DOOA) is variously described as yellow, greenish-yellow, brownish-yellow or orange. The first known use of citrine as a color name in English was in the 14th century.

C6. Let $k$ and $n$ be fixed positive integers. In the liar's guessing game, Amy chooses integers $x$ and $N$ with $1 \leq x \leq N$. She tells Ben what $N$ is, but not what $x$ is. Ben may then repeatedly ask Amy whether $x \in S$ for arbitrary sets $S$ of integers. Amy will always answer with yes or no, but she might lie. The only restriction is that she can lie at most $k$ times in a row. After he has asked as many questions as he wants, Ben must specify a set of at most $n$ positive integers. If $x$ is in this set he wins; otherwise, he loses. Prove that:
a) If $n \geq 2^{k}$ then Ben can always win.
b) For sufficiently large $k$ there exist $n \geq 1.99^{k}$ such that Ben cannot guarantee a win.

Solution. Consider an answer $A \in\{y e s, n o\}$ to a question of the kind "Is $x$ in the set $S$ ?" We say that $A$ is inconsistent with a number $i$ if $A=$ yes and $i \notin S$, or if $A=n o$ and $i \in S$. Observe that an answer inconsistent with the target number $x$ is a lie.
a) Suppose that Ben has determined a set $T$ of size $m$ that contains $x$. This is true initially with $m=N$ and $T=\{1,2, \ldots, N\}$. For $m>2^{k}$ we show how Ben can find a number $y \in T$ that is different from $x$. By performing this step repeatedly he can reduce $T$ to be of size $2^{k} \leq n$ and thus win.

Since only the size $m>2^{k}$ of $T$ is relevant, assume that $T=\left\{0,1, \ldots, 2^{k}, \ldots, m-1\right\}$. Ben begins by asking repeatedly whether $x$ is $2^{k}$. If Amy answers no $k+1$ times in a row, one of these answers is truthful, and so $x \neq 2^{k}$. Otherwise Ben stops asking about $2^{k}$ at the first answer yes. He then asks, for each $i=1, \ldots, k$, if the binary representation of $x$ has a 0 in the $i$ th digit. Regardless of what the $k$ answers are, they are all inconsistent with a certain number $y \in\left\{0,1, \ldots, 2^{k}-1\right\}$. The preceding answer yes about $2^{k}$ is also inconsistent with $y$. Hence $y \neq x$. Otherwise the last $k+1$ answers are not truthful, which is impossible.

Either way, Ben finds a number in $T$ that is different from $x$, and the claim is proven.
b) We prove that if $1<\lambda<2$ and $n=\left\lfloor(2-\lambda) \lambda^{k+1}\right\rfloor-1$ then Ben cannot guarantee a win. To complete the proof, then it suffices to take $\lambda$ such that $1.99<\lambda<2$ and $k$ large enough so that

$$
n=\left\lfloor(2-\lambda) \lambda^{k+1}\right\rfloor-1 \geq 1.99^{k}
$$

Consider the following strategy for Amy. First she chooses $N=n+1$ and $x \in\{1,2, \ldots, n+1\}$ arbitrarily. After every answer of hers Amy determines, for each $i=1,2, \ldots, n+1$, the number $m_{i}$ of consecutive answers she has given by that point that are inconsistent with $i$. To decide on her next answer, she then uses the quantity

$$
\phi=\sum_{i=1}^{n+1} \lambda^{m_{i}} .
$$

No matter what Ben's next question is, Amy chooses the answer which minimizes $\phi$.
We claim that with this strategy $\phi$ will always stay less than $\lambda^{k+1}$. Consequently no exponent $m_{i}$ in $\phi$ will ever exceed $k$, hence Amy will never give more than $k$ consecutive answers inconsistent with some $i$. In particular this applies to the target number $x$, so she will never lie more than $k$ times in a row. Thus, given the claim, Amy's strategy is legal. Since the strategy does not depend on $x$ in any way, Ben can make no deductions about $x$, and therefore he cannot guarantee a win.

It remains to show that $\phi<\lambda^{k+1}$ at all times. Initially each $m_{i}$ is 0 , so this condition holds in the beginning due to $1<\lambda<2$ and $n=\left\lfloor(2-\lambda) \lambda^{k+1}\right\rfloor-1$. Suppose that $\phi<\lambda^{k+1}$ at some point, and Ben has just asked if $x \in S$ for some set $S$. According as Amy answers yes or no, the new value of $\phi$ becomes

$$
\phi_{1}=\sum_{i \in S} 1+\sum_{i \notin S} \lambda^{m_{i}+1} \quad \text { or } \quad \phi_{2}=\sum_{i \in S} \lambda^{m_{i}+1}+\sum_{i \notin S} 1 .
$$

Since Amy chooses the option minimizing $\phi$, the new $\phi$ will equal $\min \left(\phi_{1}, \phi_{2}\right)$. Now we have

$$
\min \left(\phi_{1}, \phi_{2}\right) \leq \frac{1}{2}\left(\phi_{1}+\phi_{2}\right)=\frac{1}{2}\left(\sum_{i \in S}\left(1+\lambda^{m_{i}+1}\right)+\sum_{i \notin S}\left(\lambda^{m_{i}+1}+1\right)\right)=\frac{1}{2}(\lambda \phi+n+1) .
$$

Because $\phi<\lambda^{k+1}$, the assumptions $\lambda<2$ and $n=\left\lfloor(2-\lambda) \lambda^{k+1}\right\rfloor-1$ lead to

$$
\min \left(\phi_{1}, \phi_{2}\right)<\frac{1}{2}\left(\lambda^{k+2}+(2-\lambda) \lambda^{k+1}\right)=\lambda^{k+1} .
$$

The claim follows, which completes the solution.

Comment. Given a fixed $k$, let $f(k)$ denote the minimum value of $n$ for which Ben can guarantee a victory. The problem asks for a proof that for large $k$

$$
1.99^{k} \leq f(k) \leq 2^{k} .
$$

A computer search shows that $f(k)=2,3,4,7,11,17$ for $k=1,2,3,4,5,6$.

C7. There are given $2^{500}$ points on a circle labeled $1,2, \ldots, 2^{500}$ in some order. Prove that one can choose 100 pairwise disjoint chords joining some of these points so that the 100 sums of the pairs of numbers at the endpoints of the chosen chords are equal.

Solution. The proof is based on the following general fact.
Lemma. In a graph $G$ each vertex $v$ has degree $d_{v}$. Then $G$ contains an independent set $S$ of vertices such that $|S| \geq f(G)$ where

$$
f(G)=\sum_{v \in G} \frac{1}{d_{v}+1}
$$

Proof. Induction on $n=|G|$. The base $n=1$ is clear. For the inductive step choose a vertex $v_{0}$ in $G$ of minimum degree $d$. Delete $v_{0}$ and all of its neighbors $v_{1}, \ldots, v_{d}$ and also all edges with endpoints $v_{0}, v_{1}, \ldots, v_{d}$. This gives a new graph $G^{\prime}$. By the inductive assumption $G^{\prime}$ contains an independent set $S^{\prime}$ of vertices such that $\left|S^{\prime}\right| \geq f\left(G^{\prime}\right)$. Since no vertex in $S^{\prime}$ is a neighbor of $v_{0}$ in $G$, the set $S=S^{\prime} \cup\left\{v_{0}\right\}$ is independent in $G$.

Let $d_{v}^{\prime}$ be the degree of a vertex $v$ in $G^{\prime}$. Clearly $d_{v}^{\prime} \leq d_{v}$ for every such vertex $v$, and also $d_{v_{i}} \geq d$ for all $i=0,1, \ldots, d$ by the minimal choice of $v_{0}$. Therefore

$$
f\left(G^{\prime}\right)=\sum_{v \in G^{\prime}} \frac{1}{d_{v}^{\prime}+1} \geq \sum_{v \in G^{\prime}} \frac{1}{d_{v}+1}=f(G)-\sum_{i=0}^{d} \frac{1}{d_{v_{i}}+1} \geq f(G)-\frac{d+1}{d+1}=f(G)-1 .
$$

Hence $|S|=\left|S^{\prime}\right|+1 \geq f\left(G^{\prime}\right)+1 \geq f(G)$, and the induction is complete.
We pass on to our problem. For clarity denote $n=2^{499}$ and draw all chords determined by the given $2 n$ points. Color each chord with one of the colors $3,4, \ldots, 4 n-1$ according to the sum of the numbers at its endpoints. Chords with a common endpoint have different colors. For each color $c$ consider the following graph $G_{c}$. Its vertices are the chords of color $c$, and two chords are neighbors in $G_{c}$ if they intersect. Let $f\left(G_{c}\right)$ have the same meaning as in the lemma for all graphs $G_{c}$.

Every chord $\ell$ divides the circle into two arcs, and one of them contains $m(\ell) \leq n-1$ given points. (In particular $m(\ell)=0$ if $\ell$ joins two consecutive points.) For each $i=0,1, \ldots, n-2$ there are $2 n$ chords $\ell$ with $m(\ell)=i$. Such a chord has degree at most $i$ in the respective graph. Indeed let $A_{1}, \ldots, A_{i}$ be all points on either arc determined by a chord $\ell$ with $m(\ell)=i$ and color $c$. Every $A_{j}$ is an endpoint of at most 1 chord colored $c, j=1, \ldots, i$. Hence at most $i$ chords of color $c$ intersect $\ell$.

It follows that for each $i=0,1, \ldots, n-2$ the $2 n$ chords $\ell$ with $m(\ell)=i$ contribute at least $\frac{2 n}{i+1}$ to the sum $\sum_{c} f\left(G_{c}\right)$. Summation over $i=0,1, \ldots, n-2$ gives

$$
\sum_{c} f\left(G_{c}\right) \geq 2 n \sum_{i=1}^{n-1} \frac{1}{i}
$$

Because there are $4 n-3$ colors in all, averaging yields a color $c$ such that

$$
f\left(G_{c}\right) \geq \frac{2 n}{4 n-3} \sum_{i=1}^{n-1} \frac{1}{i}>\frac{1}{2} \sum_{i=1}^{n-1} \frac{1}{i} .
$$

By the lemma there are at least $\frac{1}{2} \sum_{i=1}^{n-1} \frac{1}{i}$ pairwise disjoint chords of color $c$, i. e. with the same sum $c$ of the pairs of numbers at their endpoints. It remains to show that $\frac{1}{2} \sum_{i=1}^{n-1} \frac{1}{i} \geq 100$ for $n=2^{499}$. Indeed we have

$$
\sum_{i=1}^{n-1} \frac{1}{i}>\sum_{i=1}^{2^{400}} \frac{1}{i}=1+\sum_{k=1}^{400} \sum_{i=2^{k-1+1}}^{2^{k}} \frac{1}{i}>1+\sum_{k=1}^{400} \frac{2^{k-1}}{2^{k}}=201>200
$$

This completes the solution.

## Geometry

G1. In the triangle $A B C$ the point $J$ is the center of the excircle opposite to $A$. This excircle is tangent to the side $B C$ at $M$, and to the lines $A B$ and $A C$ at $K$ and $L$ respectively. The lines $L M$ and $B J$ meet at $F$, and the lines $K M$ and $C J$ meet at $G$. Let $S$ be the point of intersection of the lines $A F$ and $B C$, and let $T$ be the point of intersection of the lines $A G$ and $B C$. Prove that $M$ is the midpoint of $S T$.

Solution. Let $\alpha=\angle C A B, \beta=\angle A B C$ and $\gamma=\angle B C A$. The line $A J$ is the bisector of $\angle C A B$, so $\angle J A K=\angle J A L=\frac{\alpha}{2}$. By $\angle A K J=\angle A L J=90^{\circ}$ the points $K$ and $L$ lie on the circle $\omega$ with diameter $A J$.

The triangle $K B M$ is isosceles as $B K$ and $B M$ are tangents to the excircle. Since $B J$ is the bisector of $\angle K B M$, we have $\angle M B J=90^{\circ}-\frac{\beta}{2}$ and $\angle B M K=\frac{\beta}{2}$. Likewise $\angle M C J=90^{\circ}-\frac{\gamma}{2}$ and $\angle C M L=\frac{\gamma}{2}$. Also $\angle B M F=\angle C M L$, therefore

$$
\angle L F J=\angle M B J-\angle B M F=\left(90^{\circ}-\frac{\beta}{2}\right)-\frac{\gamma}{2}=\frac{\alpha}{2}=\angle L A J .
$$

Hence $F$ lies on the circle $\omega$. (By the angle computation, $F$ and $A$ are on the same side of $B C$.) Analogously, $G$ also lies on $\omega$. Since $A J$ is a diameter of $\omega$, we obtain $\angle A F J=\angle A G J=90^{\circ}$.


The lines $A B$ and $B C$ are symmetric with respect to the external bisector $B F$. Because $A F \perp B F$ and $K M \perp B F$, the segments $S M$ and $A K$ are symmetric with respect to $B F$, hence $S M=A K$. By symmetry $T M=A L$. Since $A K$ and $A L$ are equal as tangents to the excircle, it follows that $S M=T M$, and the proof is complete.

Comment. After discovering the circle $A F K J L G$, there are many other ways to complete the solution. For instance, from the cyclic quadrilaterals $J M F S$ and $J M G T$ one can find $\angle T S J=\angle S T J=\frac{\alpha}{2}$. Another possibility is to use the fact that the lines $A S$ and $G M$ are parallel (both are perpendicular to the external angle bisector $B J$ ), so $\frac{M S}{M T}=\frac{A G}{G T}=1$.

G2. Let $A B C D$ be a cyclic quadrilateral whose diagonals $A C$ and $B D$ meet at $E$. The extensions of the sides $A D$ and $B C$ beyond $A$ and $B$ meet at $F$. Let $G$ be the point such that $E C G D$ is a parallelogram, and let $H$ be the image of $E$ under reflection in $A D$. Prove that $D, H, F, G$ are concyclic.

Solution. We show first that the triangles $F D G$ and $F B E$ are similar. Since $A B C D$ is cyclic, the triangles $E A B$ and $E D C$ are similar, as well as $F A B$ and $F C D$. The parallelogram $E C G D$ yields $G D=E C$ and $\angle C D G=\angle D C E$; also $\angle D C E=\angle D C A=\angle D B A$ by inscribed angles. Therefore

$$
\begin{gathered}
\angle F D G=\angle F D C+\angle C D G=\angle F B A+\angle A B D=\angle F B E, \\
\frac{G D}{E B}=\frac{C E}{E B}=\frac{C D}{A B}=\frac{F D}{F B} .
\end{gathered}
$$

It follows that $F D G$ and $F B E$ are similar, and so $\angle F G D=\angle F E B$.


Since $H$ is the reflection of $E$ with respect to $F D$, we conclude that

$$
\angle F H D=\angle F E D=180^{\circ}-\angle F E B=180^{\circ}-\angle F G D .
$$

This proves that $D, H, F, G$ are concyclic.
Comment. Points $E$ and $G$ are always in the half-plane determined by the line $F D$ that contains $B$ and $C$, but $H$ is always in the other half-plane. In particular, $D H F G$ is cyclic if and only if $\angle F H D+\angle F G D=180^{\circ}$.

G3. In an acute triangle $A B C$ the points $D, E$ and $F$ are the feet of the altitudes through $A$, $B$ and $C$ respectively. The incenters of the triangles $A E F$ and $B D F$ are $I_{1}$ and $I_{2}$ respectively; the circumcenters of the triangles $A C I_{1}$ and $B C I_{2}$ are $O_{1}$ and $O_{2}$ respectively. Prove that $I_{1} I_{2}$ and $O_{1} O_{2}$ are parallel.

Solution. Let $\angle C A B=\alpha, \angle A B C=\beta, \angle B C A=\gamma$. We start by showing that $A, B, I_{1}$ and $I_{2}$ are concyclic. Since $A I_{1}$ and $B I_{2}$ bisect $\angle C A B$ and $\angle A B C$, their extensions beyond $I_{1}$ and $I_{2}$ meet at the incenter $I$ of the triangle. The points $E$ and $F$ are on the circle with diameter $B C$, so $\angle A E F=\angle A B C$ and $\angle A F E=\angle A C B$. Hence the triangles $A E F$ and $A B C$ are similar with ratio of similitude $\frac{A E}{A B}=\cos \alpha$. Because $I_{1}$ and $I$ are their incenters, we obtain $I_{1} A=I A \cos \alpha$ and $I I_{1}=I A-I_{1} A=2 I A \sin ^{2} \frac{\alpha}{2}$. By symmetry $I I_{2}=2 I B \sin ^{2} \frac{\beta}{2}$. The law of sines in the triangle $A B I$ gives $I A \sin \frac{\alpha}{2}=I B \sin \frac{\beta}{2}$. Hence

$$
I I_{1} \cdot I A=2\left(I A \sin \frac{\alpha}{2}\right)^{2}=2\left(I B \sin \frac{\beta}{2}\right)^{2}=I I_{2} \cdot I B .
$$

Therefore $A, B, I_{1}$ and $I_{2}$ are concyclic, as claimed.


In addition $I I_{1} \cdot I A=I I_{2} \cdot I B$ implies that $I$ has the same power with respect to the circles $\left(A C I_{1}\right),\left(B C I_{2}\right)$ and $\left(A B I_{1} I_{2}\right)$. Then $C I$ is the radical axis of $\left(A C I_{1}\right)$ and $\left(B C I_{2}\right)$; in particular $C I$ is perpendicular to the line of centers $O_{1} O_{2}$.

Now it suffices to prove that $C I \perp I_{1} I_{2}$. Let $C I$ meet $I_{1} I_{2}$ at $Q$, then it is enough to check that $\angle I I_{1} Q+\angle I_{1} I Q=90^{\circ}$. Since $\angle I_{1} I Q$ is external for the triangle $A C I$, we have

$$
\angle I I_{1} Q+\angle I_{1} I Q=\angle I I_{1} Q+(\angle A C I+\angle C A I)=\angle I I_{1} I_{2}+\angle A C I+\angle C A I .
$$

It remains to note that $\angle I I_{1} I_{2}=\frac{\beta}{2}$ from the cyclic quadrilateral $A B I_{1} I_{2}$, and $\angle A C I=\frac{\gamma}{2}$, $\angle C A I=\frac{\alpha}{2}$. Therefore $\angle I I_{1} Q+\angle I_{1} I Q=\frac{\alpha}{2}+\frac{\beta}{2}+\frac{\gamma}{2}=90^{\circ}$, completing the proof.

Comment. It follows from the first part of the solution that the common point $I_{3} \neq C$ of the circles $\left(A C I_{1}\right)$ and $\left(B C I_{2}\right)$ is the incenter of the triangle $C D E$.

G4. Let $A B C$ be a triangle with $A B \neq A C$ and circumcenter $O$. The bisector of $\angle B A C$ intersects $B C$ at $D$. Let $E$ be the reflection of $D$ with respect to the midpoint of $B C$. The lines through $D$ and $E$ perpendicular to $B C$ intersect the lines $A O$ and $A D$ at $X$ and $Y$ respectively. Prove that the quadrilateral $B X C Y$ is cyclic.

Solution. The bisector of $\angle B A C$ and the perpendicular bisector of $B C$ meet at $P$, the midpoint of the minor arc $\widehat{B C}$ (they are different lines as $A B \neq A C$ ). In particular $O P$ is perpendicular to $B C$ and intersects it at $M$, the midpoint of $B C$.

Denote by $Y^{\prime}$ the reflexion of $Y$ with respect to $O P$. Since $\angle B Y C=\angle B Y^{\prime} C$, it suffices to prove that $B X C Y^{\prime}$ is cyclic.


We have

$$
\angle X A P=\angle O P A=\angle E Y P
$$

The first equality holds because $O A=O P$, and the second one because $E Y$ and $O P$ are both perpendicular to $B C$ and hence parallel. But $\left\{Y, Y^{\prime}\right\}$ and $\{E, D\}$ are pairs of symmetric points with respect to $O P$, it follows that $\angle E Y P=\angle D Y^{\prime} P$ and hence

$$
\angle X A P=\angle D Y^{\prime} P=\angle X Y^{\prime} P
$$

The last equation implies that $X A Y^{\prime} P$ is cyclic. By the powers of $D$ with respect to the circles $\left(X A Y^{\prime} P\right)$ and $(A B P C)$ we obtain

$$
X D \cdot D Y^{\prime}=A D \cdot D P=B D \cdot D C
$$

It follows that $B X C Y^{\prime}$ is cyclic, as desired.

G5. Let $A B C$ be a triangle with $\angle B C A=90^{\circ}$, and let $C_{0}$ be the foot of the altitude from $C$. Choose a point $X$ in the interior of the segment $C C_{0}$, and let $K, L$ be the points on the segments $A X, B X$ for which $B K=B C$ and $A L=A C$ respectively. Denote by $M$ the intersection of $A L$ and $B K$. Show that $M K=M L$.

Solution. Let $C^{\prime}$ be the reflection of $C$ in the line $A B$, and let $\omega_{1}$ and $\omega_{2}$ be the circles with centers $A$ and $B$, passing through $L$ and $K$ respectively. Since $A C^{\prime}=A C=A L$ and $B C^{\prime}=B C=B K$, both $\omega_{1}$ and $\omega_{2}$ pass through $C$ and $C^{\prime}$. By $\angle B C A=90^{\circ}, A C$ is tangent to $\omega_{2}$ at $C$, and $B C$ is tangent to $\omega_{1}$ at $C$. Let $K_{1} \neq K$ be the second intersection of $A X$ and $\omega_{2}$, and let $L_{1} \neq L$ be the second intersection of $B X$ and $\omega_{1}$.


By the powers of $X$ with respect to $\omega_{2}$ and $\omega_{1}$,

$$
X K \cdot X K_{1}=X C \cdot X C^{\prime}=X L \cdot X L_{1}
$$

so the points $K_{1}, L, K, L_{1}$ lie on a circle $\omega_{3}$.
The power of $A$ with respect to $\omega_{2}$ gives

$$
A L^{2}=A C^{2}=A K \cdot A K_{1},
$$

indicating that $A L$ is tangent to $\omega_{3}$ at $L$. Analogously, $B K$ is tangent to $\omega_{3}$ at $K$. Hence $M K$ and $M L$ are the two tangents from $M$ to $\omega_{3}$ and therefore $M K=M L$.

G6. Let $A B C$ be a triangle with circumcenter $O$ and incenter $I$. The points $D, E$ and $F$ on the sides $B C, C A$ and $A B$ respectively are such that $B D+B F=C A$ and $C D+C E=A B$. The circumcircles of the triangles $B F D$ and $C D E$ intersect at $P \neq D$. Prove that $O P=O I$.

Solution. By Miquel's theorem the circles $(A E F)=\omega_{A},(B F D)=\omega_{B}$ and $(C D E)=\omega_{C}$ have a common point, for arbitrary points $D, E$ and $F$ on $B C, C A$ and $A B$. So $\omega_{A}$ passes through the common point $P \neq D$ of $\omega_{B}$ and $\omega_{C}$.

Let $\omega_{A}, \omega_{B}$ and $\omega_{C}$ meet the bisectors $A I, B I$ and $C I$ at $A \neq A^{\prime}, B \neq B^{\prime}$ and $C \neq C^{\prime}$ respectively. The key observation is that $A^{\prime}, B^{\prime}$ and $C^{\prime}$ do not depend on the particular choice of $D, E$ and $F$, provided that $B D+B F=C A, C D+C E=A B$ and $A E+A F=B C$ hold true (the last equality follows from the other two). For a proof we need the following fact.

Lemma. Given is an angle with vertex $A$ and measure $\alpha$. A circle $\omega$ through $A$ intersects the angle bisector at $L$ and sides of the angle at $X$ and $Y$. Then $A X+A Y=2 A L \cos \frac{\alpha}{2}$.
Proof. Note that $L$ is the midpoint of arc $\widehat{X L Y}$ in $\omega$ and set $X L=Y L=u, X Y=v$. By Ptolemy's theorem $A X \cdot Y L+A Y \cdot X L=A L \cdot X Y$, which rewrites as $(A X+A Y) u=A L \cdot v$. Since $\angle L X Y=\frac{\alpha}{2}$ and $\angle X L Y=180^{\circ}-\alpha$, we have $v=2 \cos \frac{\alpha}{2} u$ by the law of sines, and the claim follows.


Apply the lemma to $\angle B A C=\alpha$ and the circle $\omega=\omega_{A}$, which intersects $A I$ at $A^{\prime}$. This gives $2 A A^{\prime} \cos \frac{\alpha}{2}=A E+A F=B C$; by symmetry analogous relations hold for $B B^{\prime}$ and $C C^{\prime}$. It follows that $A^{\prime}, B^{\prime}$ and $C^{\prime}$ are independent of the choice of $D, E$ and $F$, as stated.

We use the lemma two more times with $\angle B A C=\alpha$. Let $\omega$ be the circle with diameter $A I$. Then $X$ and $Y$ are the tangency points of the incircle of $A B C$ with $A B$ and $A C$, and hence $A X=A Y=\frac{1}{2}(A B+A C-B C)$. So the lemma yields $2 A I \cos \frac{\alpha}{2}=A B+A C-B C$. Next, if $\omega$ is the circumcircle of $A B C$ and $A I$ intersects $\omega$ at $M \neq A$ then $\{X, Y\}=\{B, C\}$, and so $2 A M \cos \frac{\alpha}{2}=A B+A C$ by the lemma. To summarize,

$$
\begin{equation*}
2 A A^{\prime} \cos \frac{\alpha}{2}=B C, \quad 2 A I \cos \frac{\alpha}{2}=A B+A C-B C, \quad 2 A M \cos \frac{\alpha}{2}=A B+A C . \tag{*}
\end{equation*}
$$

These equalities imply $A A^{\prime}+A I=A M$, hence the segments $A M$ and $I A^{\prime}$ have a common midpoint. It follows that $I$ and $A^{\prime}$ are equidistant from the circumcenter $O$. By symmetry $O I=O A^{\prime}=O B^{\prime}=O C^{\prime}$, so $I, A^{\prime}, B^{\prime}, C^{\prime}$ are on a circle centered at $O$.

To prove $O P=O I$, now it suffices to show that $I, A^{\prime}, B^{\prime}, C^{\prime}$ and $P$ are concyclic. Clearly one can assume $P \neq I, A^{\prime}, B^{\prime}, C^{\prime}$.

We use oriented angles to avoid heavy case distinction. The oriented angle between the lines $l$ and $m$ is denoted by $\angle(l, m)$. We have $\angle(l, m)=-\angle(m, l)$ and $\angle(l, m)+\angle(m, n)=\angle(l, n)$ for arbitrary lines $l, m$ and $n$. Four distinct non-collinear points $U, V, X, Y$ are concyclic if and only if $\angle(U X, V X)=\angle(U Y, V Y)$.


Suppose for the moment that $A^{\prime}, B^{\prime}, P, I$ are distinct and noncollinear; then it is enough to check the equality $\angle\left(A^{\prime} P, B^{\prime} P\right)=\angle\left(A^{\prime} I, B^{\prime} I\right)$. Because $A, F, P, A^{\prime}$ are on the circle $\omega_{A}$, we have $\angle\left(A^{\prime} P, F P\right)=\angle\left(A^{\prime} A, F A\right)=\angle\left(A^{\prime} I, A B\right)$. Likewise $\angle\left(B^{\prime} P, F P\right)=\angle\left(B^{\prime} I, A B\right)$. Therefore

$$
\angle\left(A^{\prime} P, B^{\prime} P\right)=\angle\left(A^{\prime} P, F P\right)+\angle\left(F P, B^{\prime} P\right)=\angle\left(A^{\prime} I, A B\right)-\angle\left(B^{\prime} I, A B\right)=\angle\left(A^{\prime} I, B^{\prime} I\right)
$$

Here we assumed that $P \neq F$. If $P=F$ then $P \neq D, E$ and the conclusion follows similarly (use $\angle\left(A^{\prime} F, B^{\prime} F\right)=\angle\left(A^{\prime} F, E F\right)+\angle(E F, D F)+\angle\left(D F, B^{\prime} F\right)$ and inscribed angles in $\left.\omega_{A}, \omega_{B}, \omega_{C}\right)$.

There is no loss of generality in assuming $A^{\prime}, B^{\prime}, P, I$ distinct and noncollinear. If $A B C$ is an equilateral triangle then the equalities $\left(^{*}\right)$ imply that $A^{\prime}, B^{\prime}, C^{\prime}, I, O$ and $P$ coincide, so $O P=O I$. Otherwise at most one of $A^{\prime}, B^{\prime}, C^{\prime}$ coincides with $I$. If say $C^{\prime}=I$ then $O I \perp C I$ by the previous reasoning. It follows that $A^{\prime}, B^{\prime} \neq I$ and hence $A^{\prime} \neq B^{\prime}$. Finally $A^{\prime}, B^{\prime}$ and $I$ are noncollinear because $I, A^{\prime}, B^{\prime}, C^{\prime}$ are concyclic.

Comment. The proposer remarks that the locus $\gamma$ of the points $P$ is an arc of the circle $\left(A^{\prime} B^{\prime} C^{\prime} I\right)$. The reflection $I^{\prime}$ of $I$ in $O$ belongs to $\gamma$; it is obtained by choosing $D, E$ and $F$ to be the tangency points of the three excircles with their respective sides. The rest of the circle ( $\left.A^{\prime} B^{\prime} C^{\prime} I\right)$, except $I$, can be included in $\gamma$ by letting $D, E$ and $F$ vary on the extensions of the sides and assuming signed lengths. For instance if $B$ is between $C$ and $D$ then the length $B D$ must be taken with a negative sign. The incenter $I$ corresponds to the limit case where $D$ tends to infinity.

G7. Let $A B C D$ be a convex quadrilateral with non-parallel sides $B C$ and $A D$. Assume that there is a point $E$ on the side $B C$ such that the quadrilaterals $A B E D$ and $A E C D$ are circumscribed. Prove that there is a point $F$ on the side $A D$ such that the quadrilaterals $A B C F$ and $B C D F$ are circumscribed if and only if $A B$ is parallel to $C D$.

Solution. Let $\omega_{1}$ and $\omega_{2}$ be the incircles and $O_{1}$ and $O_{2}$ the incenters of the quadrilaterals $A B E D$ and $A E C D$ respectively. A point $F$ with the stated property exists only if $\omega_{1}$ and $\omega_{2}$ are also the incircles of the quadrilaterals $A B C F$ and $B C D F$.


Let the tangents from $B$ to $\omega_{2}$ and from $C$ to $\omega_{1}$ (other than $B C$ ) meet $A D$ at $F_{1}$ and $F_{2}$ respectively. We need to prove that $F_{1}=F_{2}$ if and only if $A B \| C D$.
Lemma. The circles $\omega_{1}$ and $\omega_{2}$ with centers $O_{1}$ and $O_{2}$ are inscribed in an angle with vertex $O$. The points $P, S$ on one side of the angle and $Q, R$ on the other side are such that $\omega_{1}$ is the incircle of the triangle $P Q O$, and $\omega_{2}$ is the excircle of the triangle $R S O$ opposite to $O$. Denote $p=O O_{1} \cdot O O_{2}$. Then exactly one of the following relations holds:

$$
O P \cdot O R<p<O Q \cdot O S, \quad O P \cdot O R>p>O Q \cdot O S, \quad O P \cdot O R=p=O Q \cdot O S
$$

Proof. Denote $\angle O P O_{1}=u, \angle O Q O_{1}=v, \angle O O_{2} R=x, \angle O O_{2} S=y, \angle P O Q=2 \varphi$. Because $P O_{1}, Q O_{1}, R O_{2}, S O_{2}$ are internal or external bisectors in the triangles $P Q O$ and $R S O$, we have

$$
\begin{equation*}
u+v=x+y\left(=90^{\circ}-\varphi\right) . \tag{1}
\end{equation*}
$$



By the law of sines

$$
\frac{O P}{O O_{1}}=\frac{\sin (u+\varphi)}{\sin u} \quad \text { and } \quad \frac{O O_{2}}{O R}=\frac{\sin (x+\varphi)}{\sin x}
$$

Therefore, since $x, u$ and $\varphi$ are acute,
$O P \cdot O R \geq p \Leftrightarrow \frac{O P}{O O_{1}} \geq \frac{O O_{2}}{O R} \Leftrightarrow \sin x \sin (u+\varphi) \geq \sin u \sin (x+\varphi) \Leftrightarrow \sin (x-u) \geq 0 \Leftrightarrow x \geq u$.
Thus $O P \cdot O R \geq p$ is equivalent to $x \geq u$, with $O P \cdot O R=p$ if and only if $x=u$.
Analogously, $p \geq O Q \cdot O S$ is equivalent to $v \geq y$, with $p=O Q \cdot O S$ if and only if $v=y$. On the other hand $x \geq u$ and $v \geq y$ are equivalent by (1), with $x=u$ if and only if $v=y$. The conclusion of the lemma follows from here.

Going back to the problem, apply the lemma to the quadruples $\left\{B, E, D, F_{1}\right\},\{A, B, C, D\}$ and $\left\{A, E, C, F_{2}\right\}$. Assuming $O E \cdot O F_{1}>p$, we obtain

$$
O E \cdot O F_{1}>p \Rightarrow O B \cdot O D<p \Rightarrow O A \cdot O C>p \Rightarrow O E \cdot O F_{2}<p
$$

In other words, $O E \cdot O F_{1}>p$ implies

$$
O B \cdot O D<p<O A \cdot O C \quad \text { and } \quad O E \cdot O F_{1}>p>O E \cdot O F_{2} .
$$

Similarly, $O E \cdot O F_{1}<p$ implies

$$
O B \cdot O D>p>O A \cdot O C \quad \text { and } \quad O E \cdot O F_{1}<p<O E \cdot O F_{2} .
$$

In these cases $F_{1} \neq F_{2}$ and $O B \cdot O D \neq O A \cdot O C$, so the lines $A B$ and $C D$ are not parallel.
There remains the case $O E \cdot O F_{1}=p$. Here the lemma leads to $O B \cdot O D=p=O A \cdot O C$ and $O E \cdot O F_{1}=p=O E \cdot O F_{2}$. Therefore $F_{1}=F_{2}$ and $A B \| C D$.

Comment. The conclusion is also true if $B C$ and $A D$ are parallel. One can prove a limit case of the lemma for the configuration shown in the figure below, where $r_{1}$ and $r_{2}$ are parallel rays starting at $O^{\prime}$ and $O^{\prime \prime}$, with $O^{\prime} O^{\prime \prime} \perp r_{1}, r_{2}$ and $O$ the midpoint of $O^{\prime} O^{\prime \prime}$. Two circles with centers $O_{1}$ and $O_{2}$ are inscribed in the strip between $r_{1}$ and $r_{2}$. The lines $P Q$ and $R S$ are tangent to the circles, with $P, S$ on $r_{1}$, and $Q, R$ on $r_{2}$, so that $O, O_{1}$ are on the same side of $P Q$ and $O, O_{2}$ are on different sides of $R S$. Denote $s=O O_{1}+O O_{2}$. Then exactly one of the following relations holds:

$$
O^{\prime} P+O^{\prime \prime} R<s<O^{\prime \prime} Q+O^{\prime} S, \quad O^{\prime} P+O^{\prime \prime} R>s>O^{\prime \prime} Q+O^{\prime} S, \quad O^{\prime} P+O^{\prime \prime} R=s=O^{\prime \prime} Q+O^{\prime} S
$$



Once this is established, the proof of the original statement for $B C \| A D$ is analogous to the one in the intersecting case. One replaces products by sums of relevant segments.

G8. Let $A B C$ be a triangle with circumcircle $\omega$ and $\ell$ a line without common points with $\omega$. Denote by $P$ the foot of the perpendicular from the center of $\omega$ to $\ell$. The side-lines $B C, C A, A B$ intersect $\ell$ at the points $X, Y, Z$ different from $P$. Prove that the circumcircles of the triangles $A X P, B Y P$ and $C Z P$ have a common point different from $P$ or are mutually tangent at $P$.

Solution 1. Let $\omega_{A}, \omega_{B}, \omega_{C}$ and $\omega$ be the circumcircles of triangles $A X P, B Y P, C Z P$ and $A B C$ respectively. The strategy of the proof is to construct a point $Q$ with the same power with respect to the four circles. Then each of $P$ and $Q$ has the same power with respect to $\omega_{A}, \omega_{B}, \omega_{C}$ and hence the three circles are coaxial. In other words they have another common point $P^{\prime}$ or the three of them are tangent at $P$.

We first give a description of the point $Q$. Let $A^{\prime} \neq A$ be the second intersection of $\omega$ and $\omega_{A}$; define $B^{\prime}$ and $C^{\prime}$ analogously. We claim that $A A^{\prime}, B B^{\prime}$ and $C C^{\prime}$ have a common point. Once this claim is established, the point just constructed will be on the radical axes of the three pairs of circles $\left\{\omega, \omega_{A}\right\},\left\{\omega, \omega_{B}\right\},\left\{\omega, \omega_{C}\right\}$. Hence it will have the same power with respect to $\omega, \omega_{A}, \omega_{B}, \omega_{C}$.


We proceed to prove that $A A^{\prime}, B B^{\prime}$ and $C C^{\prime}$ intersect at one point. Let $r$ be the circumradius of triangle $A B C$. Define the points $X^{\prime}, Y^{\prime}, Z^{\prime}$ as the intersections of $A A^{\prime}, B B^{\prime}, C C^{\prime}$ with $\ell$. Observe that $X^{\prime}, Y^{\prime}, Z^{\prime}$ do exist. If $A A^{\prime}$ is parallel to $\ell$ then $\omega_{A}$ is tangent to $\ell$; hence $X=P$ which is a contradiction. Similarly, $B B^{\prime}$ and $C C^{\prime}$ are not parallel to $\ell$.

From the powers of the point $X^{\prime}$ with respect to the circles $\omega_{A}$ and $\omega$ we get

$$
X^{\prime} P \cdot\left(X^{\prime} P+P X\right)=X^{\prime} P \cdot X^{\prime} X=X^{\prime} A^{\prime} \cdot X^{\prime} A=X^{\prime} O^{2}-r^{2}
$$

hence

$$
X^{\prime} P \cdot P X=X^{\prime} O^{2}-r^{2}-X^{\prime} P^{2}=O P^{2}-r^{2}
$$

We argue analogously for the points $Y^{\prime}$ and $Z^{\prime}$, obtaining

$$
\begin{equation*}
X^{\prime} P \cdot P X=Y^{\prime} P \cdot P Y=Z^{\prime} P \cdot P Z=O P^{2}-r^{2}=k^{2} \tag{1}
\end{equation*}
$$

In these computations all segments are regarded as directed segments. We keep the same convention for the sequel.

We prove that the lines $A A^{\prime}, B B^{\prime}, C C^{\prime}$ intersect at one point by Ceva's theorem. To avoid distracting remarks we interpret everything projectively, i. e. whenever two lines are parallel they meet at a point on the line at infinity.

Let $U, V, W$ be the intersections of $A A^{\prime}, B B^{\prime}, C C^{\prime}$ with $B C, C A, A B$ respectively. The idea is that although it is difficult to calculate the ratio $\frac{B U}{C U}$, it is easier to deal with the cross-ratio $\frac{B U}{C U} / \frac{B X}{C X}$ because we can send it to the line $\ell$. With this in mind we apply Menelaus' theorem to the triangle $A B C$ and obtain $\frac{B X}{C X} \cdot \frac{C Y}{A Y} \cdot \frac{A Z}{B Z}=1$. Hence Ceva's ratio can be expressed as

$$
\frac{B U}{C U} \cdot \frac{C V}{A V} \cdot \frac{A W}{B W}=\frac{B U}{C U} / \frac{B X}{C X} \cdot \frac{C V}{A V} / \frac{C Y}{A Y} \cdot \frac{A W}{B W} / \frac{A Z}{B Z}
$$



Project the line $B C$ to $\ell$ from $A$. The cross-ratio between $B C$ and $U X$ equals the cross-ratio between $Z Y$ and $X^{\prime} X$. Repeating the same argument with the lines $C A$ and $A B$ gives

$$
\frac{B U}{C U} \cdot \frac{C V}{A V} \cdot \frac{A W}{B W}=\frac{Z X^{\prime}}{Y X^{\prime}} / \frac{Z X}{Y X} \cdot \frac{X Y^{\prime}}{Z Y^{\prime}} / \frac{X Y}{Z Y} \cdot \frac{Y Z^{\prime}}{X Z^{\prime}} / \frac{Y Z}{X Z}
$$

and hence

$$
\frac{B U}{C U} \cdot \frac{C V}{A V} \cdot \frac{A W}{B W}=(-1) \cdot \frac{Z X^{\prime}}{Y X^{\prime}} \cdot \frac{X Y^{\prime}}{Z Y^{\prime}} \cdot \frac{Y Z^{\prime}}{X Z^{\prime}}
$$

The equations (1) reduce the problem to a straightforward computation on the line $\ell$. For instance, the transformation $t \mapsto-k^{2} / t$ preserves cross-ratio and interchanges the points $X, Y, Z$ with the points $X^{\prime}, Y^{\prime}, Z^{\prime}$. Then

$$
\frac{B U}{C U} \cdot \frac{C V}{A V} \cdot \frac{A W}{B W}=(-1) \cdot \frac{Z X^{\prime}}{Y X^{\prime}} / \frac{Z Z^{\prime}}{Y Z^{\prime}} \cdot \frac{X Y^{\prime}}{Z Y^{\prime}} / \frac{X Z^{\prime}}{Z Z^{\prime}}=-1 .
$$

We proved that Ceva's ratio equals -1 , so $A A^{\prime}, B B^{\prime}, C C^{\prime}$ intersect at one point $Q$.

Comment 1. There is a nice projective argument to prove that $A X^{\prime}, B Y^{\prime}, C Z^{\prime}$ intersect at one point. Suppose that $\ell$ and $\omega$ intersect at a pair of complex conjugate points $D$ and $E$. Consider a projective transformation that takes $D$ and $E$ to $[i ; 1,0]$ and $[-i, 1,0]$. Then $\ell$ is the line at infinity, and $\omega$ is a conic through the special points $[i ; 1,0]$ and $[-i, 1,0]$, hence it is a circle. So one can assume that $A X, B Y, C Z$ are parallel to $B C, C A, A B$. The involution on $\ell$ taking $X, Y, Z$ to $X^{\prime}, Y^{\prime}, Z^{\prime}$ and leaving $D, E$ fixed is the involution changing each direction to its perpendicular one. Hence $A X, B Y, C Z$ are also perpendicular to $A X^{\prime}, B Y^{\prime}, C Z^{\prime}$.

It follows from the above that $A X^{\prime}, B Y^{\prime}, C Z^{\prime}$ intersect at the orthocenter of triangle $A B C$.
Comment 2. The restriction that the line $\ell$ does not intersect the circumcricle $\omega$ is unnecessary. The proof above works in general. In case $\ell$ intersects $\omega$ at $D$ and $E$ point $P$ is the midpoint of $D E$, and some equations can be interpreted differently. For instance

$$
X^{\prime} P \cdot X^{\prime} X=X^{\prime} A^{\prime} \cdot X^{\prime} A=X^{\prime} D \cdot X^{\prime} E,
$$

and hence the pairs $X^{\prime} X$ and $D E$ are harmonic conjugates. This means that $X^{\prime}, Y^{\prime}, Z^{\prime}$ are the harmonic conjugates of $X, Y, Z$ with respect to the segment $D E$.

Solution 2. First we prove that there is an inversion in space that takes $\ell$ and $\omega$ to parallel circles on a sphere. Let $Q R$ be the diameter of $\omega$ whose extension beyond $Q$ passes through $P$. Let $\Pi$ be the plane carrying our objects. In space, choose a point $O$ such that the line $Q O$ is perpendicular to $\Pi$ and $\angle P O R=90^{\circ}$, and apply an inversion with pole $O$ (the radius of the inversion does not matter). For any object $\mathcal{T}$ denote by $\mathcal{T}^{\prime}$ the image of $\mathcal{T}$ under this inversion.

The inversion takes the plane $\Pi$ to a sphere $\Pi^{\prime}$. The lines in $\Pi$ are taken to circles through $O$, and the circles in $\Pi$ also are taken to circles on $\Pi^{\prime}$.


Since the line $\ell$ and the circle $\omega$ are perpendicular to the plane $O P Q$, the circles $\ell^{\prime}$ and $\omega^{\prime}$ also are perpendicular to this plane. Hence, the planes of the circles $\ell^{\prime}$ and $\omega^{\prime}$ are parallel.

Now consider the circles $A^{\prime} X^{\prime} P^{\prime}, B^{\prime} Y^{\prime} P^{\prime}$ and $C^{\prime} Z^{\prime} P^{\prime}$. We want to prove that either they have a common point (on $\Pi^{\prime}$ ), different from $P^{\prime}$, or they are tangent to each other.


The point $X^{\prime}$ is the second intersection of the circles $B^{\prime} C^{\prime} O$ and $\ell^{\prime}$, other than $O$. Hence, the lines $O X^{\prime}$ and $B^{\prime} C^{\prime}$ are coplanar. Moreover, they lie in the parallel planes of $\ell^{\prime}$ and $\omega^{\prime}$. Therefore, $O X^{\prime}$ and $B^{\prime} C^{\prime}$ are parallel. Analogously, $O Y^{\prime}$ and $O Z^{\prime}$ are parallel to $A^{\prime} C^{\prime}$ and $A^{\prime} B^{\prime}$.

Let $A_{1}$ be the second intersection of the circles $A^{\prime} X^{\prime} P^{\prime}$ and $\omega^{\prime}$, other than $A^{\prime}$. The segments $A^{\prime} A_{1}$ and $P^{\prime} X^{\prime}$ are coplanar, and therefore parallel. Now we know that $B^{\prime} C^{\prime}$ and $A^{\prime} A_{1}$ are parallel to $O X^{\prime}$ and $X^{\prime} P^{\prime}$ respectively, but these two segments are perpendicular because $O P^{\prime}$ is a diameter in $\ell^{\prime}$. We found that $A^{\prime} A_{1}$ and $B^{\prime} C^{\prime}$ are perpendicular, hence $A^{\prime} A_{1}$ is the altitude in the triangle $A^{\prime} B^{\prime} C^{\prime}$, starting from $A$.

Analogously, let $B_{1}$ and $C_{1}$ be the second intersections of $\omega^{\prime}$ with the circles $B^{\prime} P^{\prime} Y^{\prime}$ and $C^{\prime} P^{\prime} Z^{\prime}$, other than $B^{\prime}$ and $C^{\prime}$ respectively. Then $B^{\prime} B_{1}$ and $C^{\prime} C_{1}$ are the other two altitudes in the triangle $A^{\prime} B^{\prime} C^{\prime}$.

Let $H$ be the orthocenter of the triangle $A^{\prime} B^{\prime} C^{\prime}$. Let $W$ be the second intersection of the line $P^{\prime} H$ with the sphere $\Pi^{\prime}$, other than $P^{\prime}$. The point $W$ lies on the sphere $\Pi^{\prime}$, in the plane of the circle $A^{\prime} P^{\prime} X^{\prime}$, so $W$ lies on the circle $A^{\prime} P^{\prime} X^{\prime}$. Similarly, $W$ lies on the circles $B^{\prime} P^{\prime} Y^{\prime}$ and $C^{\prime} P^{\prime} Z^{\prime}$ as well; indeed $W$ is the second common point of the three circles.

If the line $P^{\prime} H$ is tangent to the sphere then $W$ coincides with $P^{\prime}$, and $P^{\prime} H$ is the common tangent of the three circles.

## Number Theory

N1. Call admissible a set $A$ of integers that has the following property:

$$
\text { If } x, y \in A \text { (possibly } x=y \text { ) then } x^{2}+k x y+y^{2} \in A \text { for every integer } k \text {. }
$$

Determine all pairs $m, n$ of nonzero integers such that the only admissible set containing both $m$ and $n$ is the set of all integers.

Solution. A pair of integers $m, n$ fulfills the condition if and only if $\operatorname{gcd}(m, n)=1$. Suppose that $\operatorname{gcd}(m, n)=d>1$. The set

$$
A=\{\ldots,-2 d,-d, 0, d, 2 d, \ldots\}
$$

is admissible, because if $d$ divides $x$ and $y$ then it divides $x^{2}+k x y+y^{2}$ for every integer $k$. Also $m, n \in A$ and $A \neq \mathbb{Z}$.

Now let $\operatorname{gcd}(m, n)=1$, and let $A$ be an admissible set containing $m$ and $n$. We use the following observations to prove that $A=\mathbb{Z}$ :
(i) $k x^{2} \in A$ for every $x \in A$ and every integer $k$.
(ii) $(x+y)^{2} \in A$ for all $x, y \in A$.

To justify (i) let $y=x$ in the definition of an admissible set; to justify (ii) let $k=2$.
Since $\operatorname{gcd}(m, n)=1$, we also have $\operatorname{gcd}\left(m^{2}, n^{2}\right)=1$. Hence one can find integers $a, b$ such that $a m^{2}+b n^{2}=1$. It follows from (i) that $a m^{2} \in A$ and $b n^{2} \in A$. Now we deduce from (ii) that $1=\left(a m^{2}+b n^{2}\right)^{2} \in A$. But if $1 \in A$ then (i) implies $k \in A$ for every integer $k$.

N2. Find all triples $(x, y, z)$ of positive integers such that $x \leq y \leq z$ and

$$
x^{3}\left(y^{3}+z^{3}\right)=2012(x y z+2) .
$$

Solution. First note that $x$ divides $2012 \cdot 2=2^{3} \cdot 503$. If $503 \mid x$ then the right-hand side of the equation is divisible by $503^{3}$, and it follows that $503^{2} \mid x y z+2$. This is false as $503 \mid x$. Hence $x=2^{m}$ with $m \in\{0,1,2,3\}$. If $m \geq 2$ then $2^{6} \mid 2012(x y z+2)$. However the highest powers of 2 dividing 2012 and $x y z+2=2^{m} y z+2$ are $2^{2}$ and $2^{1}$ respectively. So $x=1$ or $x=2$, yielding the two equations

$$
y^{3}+z^{3}=2012(y z+2), \quad \text { and } \quad y^{3}+z^{3}=503(y z+1) .
$$

In both cases the prime $503=3 \cdot 167+2$ divides $y^{3}+z^{3}$. We claim that $503 \mid y+z$. This is clear if $503 \mid y$, so let $503 \nmid y$ and $503 \nmid z$. Then $y^{502} \equiv z^{502}(\bmod 503)$ by Fermat's little theorem. On the other hand $y^{3} \equiv-z^{3}(\bmod 503)$ implies $y^{3 \cdot 167} \equiv-z^{3 \cdot 167}(\bmod 503)$, i. e. $y^{501} \equiv-z^{501}(\bmod 503)$. It follows that $y \equiv-z(\bmod 503)$ as claimed.

Therefore $y+z=503 k$ with $k \geq 1$. In view of $y^{3}+z^{3}=(y+z)\left((y-z)^{2}+y z\right)$ the two equations take the form

$$
\begin{align*}
& k(y-z)^{2}+(k-4) y z=8,  \tag{1}\\
& k(y-z)^{2}+(k-1) y z=1 . \tag{2}
\end{align*}
$$

In (1) we have $(k-4) y z \leq 8$, which implies $k \leq 4$. Indeed if $k>4$ then $1 \leq(k-4) y z \leq 8$, so that $y \leq 8$ and $z \leq 8$. This is impossible as $y+z=503 k \geq 503$. Note next that $y^{3}+z^{3}$ is even in the first equation. Hence $y+z=503 k$ is even too, meaning that $k$ is even. Thus $k=2$ or $k=4$. Clearly (1) has no integer solutions for $k=4$. If $k=2$ then (1) takes the form $(y+z)^{2}-5 y z=4$. Since $y+z=503 k=503 \cdot 2$, this leads to $5 y z=503^{2} \cdot 2^{2}-4$. However $503^{2} \cdot 2^{2}-4$ is not a multiple of 5 . Therefore (1) has no integer solutions.

Equation (2) implies $0 \leq(k-1) y z \leq 1$, so that $k=1$ or $k=2$. Also $0 \leq k(y-z)^{2} \leq 1$, hence $k=2$ only if $y=z$. However then $y=z=1$, which is false in view of $y+z \geq 503$. Therefore $k=1$ and (2) takes the form $(y-z)^{2}=1$, yielding $z-y=|y-z|=1$. Combined with $k=1$ and $y+z=503 k$, this leads to $y=251, z=252$.

In summary the triple $(2,251,252)$ is the only solution.

N3. Determine all integers $m \geq 2$ such that every $n$ with $\frac{m}{3} \leq n \leq \frac{m}{2}$ divides the binomial coefficient $\binom{n}{m-2 n}$.

Solution. The integers in question are all prime numbers.
First we check that all primes satisfy the condition, and even a stronger one. Namely, if $p$ is a prime then every $n$ with $1 \leq n \leq \frac{p}{2}$ divides $\binom{n}{p-2 n}$. This is true for $p=2$ where $n=1$ is the only possibility. For an odd prime $p$ take $n \in\left[1, \frac{p}{2}\right]$ and consider the following identity of binomial coefficients:

$$
(p-2 n) \cdot\binom{n}{p-2 n}=n \cdot\binom{n-1}{p-2 n-1} .
$$

Since $p \geq 2 n$ and $p$ is odd, all factors are non-zero. If $d=\operatorname{gcd}(p-2 n, n)$ then $d$ divides $p$, but $d \leq n<p$ and hence $d=1$. It follows that $p-2 n$ and $n$ are relatively prime, and so the factor $n$ in the right-hand side divides the binomial coefficient $\binom{n}{p-2 n}$.

Next we show that no composite number $m$ has the stated property. Consider two cases.

- If $m=2 k$ with $k>1$, pick $n=k$. Then $\frac{m}{3} \leq n \leq \frac{m}{2}$ but $\binom{n}{m-2 n}=\binom{k}{0}=1$ is not divisible by $k>1$.
- If $m$ is odd then there exist an odd prime $p$ and an integer $k \geq 1$ with $m=p(2 k+1)$. Pick $n=p k$, then $\frac{m}{3} \leq n \leq \frac{m}{2}$ by $k \geq 1$. However

$$
\frac{1}{n}\binom{n}{m-2 n}=\frac{1}{p k}\binom{p k}{p}=\frac{(p k-1)(p k-2) \cdots(p k-(p-1))}{p!}
$$

is not an integer, because $p$ divides the denominator but not the numerator.

N4. An integer $a$ is called friendly if the equation $\left(m^{2}+n\right)\left(n^{2}+m\right)=a(m-n)^{3}$ has a solution over the positive integers.
a) Prove that there are at least 500 friendly integers in the set $\{1,2, \ldots, 2012\}$.
b) Decide whether $a=2$ is friendly.

Solution. a) Every $a$ of the form $a=4 k-3$ with $k \geq 2$ is friendly. Indeed the numbers $m=2 k-1>0$ and $n=k-1>0$ satisfy the given equation with $a=4 k-3$ :

$$
\left(m^{2}+n\right)\left(n^{2}+m\right)=\left((2 k-1)^{2}+(k-1)\right)\left((k-1)^{2}+(2 k-1)\right)=(4 k-3) k^{3}=a(m-n)^{3} .
$$

Hence $5,9, \ldots, 2009$ are friendly and so $\{1,2, \ldots, 2012\}$ contains at least 502 friendly numbers.
b) We show that $a=2$ is not friendly. Consider the equation with $a=2$ and rewrite its left-hand side as a difference of squares:

$$
\frac{1}{4}\left(\left(m^{2}+n+n^{2}+m\right)^{2}-\left(m^{2}+n-n^{2}-m\right)^{2}\right)=2(m-n)^{3} .
$$

Since $m^{2}+n-n^{2}-m=(m-n)(m+n-1)$, we can further reformulate the equation as

$$
\left(m^{2}+n+n^{2}+m\right)^{2}=(m-n)^{2}\left(8(m-n)+(m+n-1)^{2}\right) .
$$

It follows that $8(m-n)+(m+n-1)^{2}$ is a perfect square. Clearly $m>n$, hence there is an integer $s \geq 1$ such that

$$
(m+n-1+2 s)^{2}=8(m-n)+(m+n-1)^{2} .
$$

Subtracting the squares gives $s(m+n-1+s)=2(m-n)$. Since $m+n-1+s>m-n$, we conclude that $s<2$. Therefore the only possibility is $s=1$ and $m=3 n$. However then the left-hand side of the given equation (with $a=2$ ) is greater than $m^{3}=27 n^{3}$, whereas its right-hand side equals $16 n^{3}$. The contradiction proves that $a=2$ is not friendly.

Comment. A computer search shows that there are 561 friendly numbers in $\{1,2, \ldots, 2012\}$.

N5. For a nonnegative integer $n$ define $\operatorname{rad}(n)=1$ if $n=0$ or $n=1$, and $\operatorname{rad}(n)=p_{1} p_{2} \cdots p_{k}$ where $p_{1}<p_{2}<\cdots<p_{k}$ are all prime factors of $n$. Find all polynomials $f(x)$ with nonnegative integer coefficients such that $\operatorname{rad}(f(n))$ divides $\operatorname{rad}\left(f\left(n^{\operatorname{rad}(n)}\right)\right)$ for every nonnegative integer $n$.

Solution 1. We are going to prove that $f(x)=a x^{m}$ for some nonnegative integers $a$ and $m$. If $f(x)$ is the zero polynomial we are done, so assume that $f(x)$ has at least one positive coefficient. In particular $f(1)>0$.

Let $p$ be a prime number. The condition is that $f(n) \equiv 0(\bmod p)$ implies

$$
\begin{equation*}
f\left(n^{\operatorname{rad}(n)}\right) \equiv 0 \quad(\bmod p) \tag{1}
\end{equation*}
$$

Since $\operatorname{rad}\left(n^{\operatorname{rad}(n)^{k}}\right)=\operatorname{rad}(n)$ for all $k$, repeated applications of the preceding implication show that if $p$ divides $f(n)$ then

$$
f\left(n^{\operatorname{rad}(n)^{k}}\right) \equiv 0 \quad(\bmod p) \quad \text { for all } k .
$$

The idea is to construct a prime $p$ and a positive integer $n$ such that $p-1$ divides $n$ and $p$ divides $f(n)$. In this case, for $k$ large enough $p-1$ divides $\operatorname{rad}(n)^{k}$. Hence if $(p, n)=1$ then $n^{\operatorname{rad}(n)^{k}} \equiv 1(\bmod p)$ by Fermat's little theorem, so that

$$
\begin{equation*}
f(1) \equiv f\left(n^{\operatorname{rad}(n)^{k}}\right) \equiv 0 \quad(\bmod p) \tag{2}
\end{equation*}
$$

Suppose that $f(x)=g(x) x^{m}$ with $g(0) \neq 0$. Let $t$ be a positive integer, $p$ any prime factor of $g(-t)$ and $n=(p-1) t$. So $p-1$ divides $n$ and $f(n)=f((p-1) t) \equiv f(-t) \equiv 0(\bmod p)$, hence either $(p, n)>1$ or $(2)$ holds. If $(p,(p-1) t)>1$ then $p$ divides $t$ and $g(0) \equiv g(-t) \equiv 0(\bmod p)$, meaning that $p$ divides $g(0)$.

In conclusion we proved that each prime factor of $g(-t)$ divides $g(0) f(1) \neq 0$, and thus the set of prime factors of $g(-t)$ when $t$ ranges through the positive integers is finite. This is known to imply that $g(x)$ is a constant polynomial, and so $f(x)=a x^{m}$.

Solution 2. Let $f(x)$ be a polynomial with integer coefficients (not necessarily nonnegative) such that $\operatorname{rad}(f(n))$ divides $\operatorname{rad}\left(f\left(n^{\operatorname{rad}(n)}\right)\right)$ for any nonnegative integer $n$. We give a complete description of all polynomials with this property. More precisely, we claim that if $f(x)$ is such a polynomial and $\xi$ is a root of $f(x)$ then so is $\xi^{d}$ for every positive integer $d$.

Therefore each root of $f(x)$ is zero or a root of unity. In particular, if a root of unity $\xi$ is a root of $f(x)$ then $1=\xi^{d}$ is a root too (for some positive integer $d$ ). In the original problem $f(x)$ has nonnegative coefficients. Then either $f(x)$ is the zero polynomial or $f(1)>0$ and $\xi=0$ is the only possible root. In either case $f(x)=a x^{m}$ with $a$ and $m$ nonnegative integers.

To prove the claim let $\xi$ be a root of $f(x)$, and let $g(x)$ be an irreducible factor of $f(x)$ such that $g(\xi)=0$. If 0 or 1 are roots of $g(x)$ then either $\xi=0$ or $\xi=1$ (because $g(x)$ is irreducible) and we are done. So assume that $g(0), g(1) \neq 0$. By decomposing $d$ as a product of prime numbers, it is enough to consider the case $d=p$ prime. We argue for $p=2$. Since $\operatorname{rad}\left(2^{k}\right)=2$ for every $k$, we have

$$
\operatorname{rad}\left(f\left(2^{k}\right)\right) \mid \operatorname{rad}\left(f\left(2^{2 k}\right)\right)
$$

Now we prove that $g(x)$ divides $f\left(x^{2}\right)$. Suppose that this is not the case. Then, since $g(x)$ is irreducible, there are integer-coefficient polynomials $a(x), b(x)$ and an integer $N$ such that

$$
\begin{equation*}
a(x) g(x)+b(x) f\left(x^{2}\right)=N \tag{3}
\end{equation*}
$$

Each prime factor $p$ of $g\left(2^{k}\right)$ divides $f\left(2^{k}\right)$, so by $\operatorname{rad}\left(f\left(2^{k}\right)\right) \mid \operatorname{rad}\left(f\left(2^{2 k}\right)\right)$ it also divides $f\left(2^{2 k}\right)$. From the equation above with $x=2^{k}$ it follows that $p$ divides $N$.

In summary, each prime divisor of $g\left(2^{k}\right)$ divides $N$, for all $k \geq 0$. Let $p_{1}, \ldots, p_{n}$ be the odd primes dividing $N$, and suppose that

$$
g(1)=2^{\alpha} p_{1}^{\alpha_{1}} \cdots p_{n}^{\alpha_{n}} .
$$

If $k$ is divisible by $\varphi\left(p_{1}^{\alpha_{1}+1} \cdots p_{n}^{\alpha_{n}+1}\right)$ then

$$
2^{k} \equiv 1 \quad\left(\bmod p_{1}^{\alpha_{1}+1} \cdots p_{n}^{\alpha_{n}+1}\right)
$$

yielding

$$
g\left(2^{k}\right) \equiv g(1) \quad\left(\bmod p_{1}^{\alpha_{1}+1} \cdots p_{n}^{\alpha_{n}+1}\right)
$$

It follows that for each $i$ the maximal power of $p_{i}$ dividing $g\left(2^{k}\right)$ and $g(1)$ is the same, namely $p_{i}^{\alpha_{i}}$. On the other hand, for large enough $k$, the maximal power of 2 dividing $g\left(2^{k}\right)$ and $g(0) \neq 0$ is the same. From the above, for $k$ divisible by $\varphi\left(p_{1}^{\alpha_{1}+1} \cdots p_{n}^{\alpha_{n}+1}\right)$ and large enough, we obtain that $g\left(2^{k}\right)$ divides $g(0) \cdot g(1)$. This is impossible because $g(0), g(1) \neq 0$ are fixed and $g\left(2^{k}\right)$ is arbitrarily large.

In conclusion, $g(x)$ divides $f\left(x^{2}\right)$. Recall that $\xi$ is a root of $f(x)$ such that $g(\xi)=0$; then $f\left(\xi^{2}\right)=0$, i. e. $\xi^{2}$ is a root of $f(x)$.

Likewise if $\xi$ is a root of $f(x)$ and $p$ an arbitrary prime then $\xi^{p}$ is a root too. The argument is completely analogous, in the proof above just replace 2 by $p$ and "odd prime" by "prime different from $p$."

Comment. The claim in the second solution can be proved by varying $n(\bmod p)$ in (1). For instance, we obtain

$$
f\left(n^{r a d(n+p k)}\right) \equiv 0 \quad(\bmod p)
$$

for every positive integer $k$. One can prove that if $(n, p)=1$ then $\operatorname{rad}(n+p k)$ runs through all residue classes $r(\bmod p-1)$ with $(r, p-1)$ squarefree. Hence if $f(n) \equiv 0(\bmod p)$ then $f\left(n^{r}\right) \equiv 0(\bmod p)$ for all integers $r$. This implies the claim by an argument leading to the identity (3).

N6. Let $x$ and $y$ be positive integers. If $x^{2^{n}}-1$ is divisible by $2^{n} y+1$ for every positive integer $n$, prove that $x=1$.

Solution. First we prove the following fact: For every positive integer $y$ there exist infinitely many primes $p \equiv 3(\bmod 4)$ such that $p$ divides some number of the form $2^{n} y+1$.

Clearly it is enough to consider the case $y$ odd. Let

$$
2 y+1=p_{1}^{e_{1}} \cdots p_{r}^{e_{r}}
$$

be the prime factorization of $2 y+1$. Suppose on the contrary that there are finitely many primes $p_{r+1}, \ldots, p_{r+s} \equiv 3(\bmod 4)$ that divide some number of the form $2^{n} y+1$ but do not divide $2 y+1$.

We want to find an $n$ such that $p_{i}^{e_{i}} \| 2^{n} y+1$ for $1 \leq i \leq r$ and $p_{i} \nmid 2^{n} y+1$ for $r+1 \leq i \leq r+s$. For this it suffices to take

$$
n=1+\varphi\left(p_{1}^{e_{1}+1} \cdots p_{r}^{e_{r}+1} p_{r+1}^{1} \cdots p_{r+s}^{1}\right)
$$

because then

$$
2^{n} y+1 \equiv 2 y+1 \quad\left(\bmod p_{1}^{e_{1}+1} \cdots p_{r}^{e_{r}+1} p_{r+1}^{1} \cdots p_{r+s}^{1}\right)
$$

The last congruence means that $p_{1}^{e_{1}}, \ldots, p_{r}^{e_{r}}$ divide exactly $2^{n} y+1$ and no prime $p_{r+1}, \ldots, p_{r+s}$ divides $2^{n} y+1$. It follows that the prime factorization of $2^{n} y+1$ consists of the prime powers $p_{1}^{e_{1}}, \ldots, p_{r}^{e_{r}}$ and powers of primes $\equiv 1(\bmod 4)$. Because $y$ is odd, we obtain

$$
2^{n} y+1 \equiv p_{1}^{e_{1}} \cdots p_{r}^{e_{r}} \equiv 2 y+1 \equiv 3 \quad(\bmod 4)
$$

This is a contradiction since $n>1$, and so $2^{n} y+1 \equiv 1(\bmod 4)$.
Now we proceed to the problem. If $p$ is a prime divisor of $2^{n} y+1$ the problem statement implies that $x^{d} \equiv 1(\bmod p)$ for $d=2^{n}$. By Fermat's little theorem the same congruence holds for $d=p-1$, so it must also hold for $d=\left(2^{n}, p-1\right)$. For $p \equiv 3(\bmod 4)$ we have $\left(2^{n}, p-1\right)=2$, therefore in this case $x^{2} \equiv 1(\bmod p)$.

In summary, we proved that every prime $p \equiv 3(\bmod 4)$ that divides some number of the form $2^{n} y+1$ also divides $x^{2}-1$. This is possible only if $x=1$, otherwise by the above $x^{2}-1$ would be a positive integer with infinitely many prime factors.

Comment. For each $x$ and each odd prime $p$ the maximal power of $p$ dividing $x^{2^{n}}-1$ for some $n$ is bounded and hence the same must be true for the numbers $2^{n} y+1$. We infer that $p^{2}$ divides $2^{p-1}-1$ for each prime divisor $p$ of $2^{n} y+1$. However trying to reach a contradiction with this conclusion alone seems hopeless, since it is not even known if there are infinitely many primes $p$ without this property.

N7. Find all $n \in \mathbb{N}$ for which there exist nonnegative integers $a_{1}, a_{2}, \ldots, a_{n}$ such that

$$
\frac{1}{2^{a_{1}}}+\frac{1}{2^{a_{2}}}+\cdots+\frac{1}{2^{a_{n}}}=\frac{1}{3^{a_{1}}}+\frac{2}{3^{a_{2}}}+\cdots+\frac{n}{3^{a_{n}}}=1 .
$$

Solution. Such numbers $a_{1}, a_{2}, \ldots, a_{n}$ exist if and only if $n \equiv 1(\bmod 4)$ or $n \equiv 2(\bmod 4)$.
Let $\sum_{k=1}^{n} \frac{k}{3^{a_{k}}}=1$ with $a_{1}, a_{2}, \ldots, a_{n}$ nonnegative integers. Then $1 \cdot x_{1}+2 \cdot x_{2}+\cdots+n \cdot x_{n}=3^{a}$ with $x_{1}, \ldots, x_{n}$ powers of 3 and $a \geq 0$. The right-hand side is odd, and the left-hand side has the same parity as $1+2+\cdots+n$. Hence the latter sum is odd, which implies $n \equiv 1,2(\bmod 4)$. Now we prove the converse.

Call feasible a sequence $b_{1}, b_{2}, \ldots, b_{n}$ if there are nonnegative integers $a_{1}, a_{2}, \ldots, a_{n}$ such that

$$
\frac{1}{2^{a_{1}}}+\frac{1}{2^{a_{2}}}+\cdots+\frac{1}{2^{a_{n}}}=\frac{b_{1}}{3^{a_{1}}}+\frac{b_{2}}{3^{a_{2}}}+\cdots+\frac{b_{n}}{3^{a_{n}}}=1 .
$$

Let $b_{k}$ be a term of a feasible sequence $b_{1}, b_{2}, \ldots, b_{n}$ with exponents $a_{1}, a_{2}, \ldots, a_{n}$ like above, and let $u, v$ be nonnegative integers with sum $3 b_{k}$. Observe that

$$
\frac{1}{2^{a_{k}+1}}+\frac{1}{2^{a_{k}+1}}=\frac{1}{2^{a_{k}}} \quad \text { and } \quad \frac{u}{3^{a_{k}+1}}+\frac{v}{3^{a_{k}+1}}=\frac{b_{k}}{3^{a_{k}}} .
$$

It follows that the sequence $b_{1}, \ldots, b_{k-1}, u, v, b_{k+1}, \ldots, b_{n}$ is feasible. The exponents $a_{i}$ are the same for the unchanged terms $b_{i}, i \neq k$; the new terms $u, v$ have exponents $a_{k}+1$.

We state the conclusion in reverse. If two terms $u, v$ of a sequence are replaced by one term $\frac{u+v}{3}$ and the obtained sequence is feasible, then the original sequence is feasible too. Denote by $\alpha_{n}$ the sequence $1,2, \ldots, n$. To show that $\alpha_{n}$ is feasible for $n \equiv 1,2(\bmod 4)$, we transform it by $n-1$ replacements $\{u, v\} \mapsto \frac{u+v}{3}$ to the one-term sequence $\alpha_{1}$. The latter is feasible, with $a_{1}=0$. Note that if $m$ and $2 m$ are terms of a sequence then $\{m, 2 m\} \mapsto m$, so $2 m$ can be ignored if necessary.

Let $n \geq 16$. We prove that $\alpha_{n}$ can be reduced to $\alpha_{n-12}$ by 12 operations. Write $n=12 k+r$ where $k \geq 1$ and $0 \leq r \leq 11$. If $0 \leq r \leq 5$ then the last 12 terms of $\alpha_{n}$ can be partitioned into 2 singletons $\{12 k-6\},\{12 k\}$ and the following 5 pairs:

$$
\{12 k-6-i, 12 k-6+i\}, i=1, \ldots, 5-r ; \quad\{12 k-j, 12 k+j\}, j=1, \ldots, r .
$$

(There is only one kind of pairs if $r \in\{0,5\}$.) One can ignore $12 k-6$ and $12 k$ since $\alpha_{n}$ contains $6 k-3$ and $6 k$. Furthermore the 5 operations $\{12 k-6-i, 12 k-6+i\} \mapsto 8 k-4$ and $\{12 k-j, 12 k+j\} \mapsto 8 k$ remove the 10 terms in the pairs and bring in 5 new terms equal to $8 k-4$ or $8 k$. All of these can be ignored too as $4 k-2$ and $4 k$ are still present in the sequence. Indeed $4 k \leq n-12$ is equivalent to $8 k \geq 12-r$, which is true for $r \in\{4,5\}$. And if $r \in\{0,1,2,3\}$ then $n \geq 16$ implies $k \geq 2$, so $8 k \geq 12-r$ also holds. Thus $\alpha_{n}$ reduces to $\alpha_{n-12}$.

The case $6 \leq r \leq 11$ is analogous. Consider the singletons $\{12 k\},\{12 k+6\}$ and the 5 pairs

$$
\{12 k-i, 12 k+i\}, i=1, \ldots, 11-r ; \quad\{12 k+6-j, 12 k+6+j\}, j=1, \ldots, r-6
$$

Ignore the singletons like before, then remove the pairs via operations $\{12 k-i, 12 k+i\} \mapsto 8 k$ and $\{12 k+6-j, 12 k+6+j\} \mapsto 8 k+4$. The 5 newly-appeared terms $8 k$ and $8 k+4$ can be ignored too since $4 k+2 \leq n-12$ (this follows from $k \geq 1$ and $r \geq 6$ ). We obtain $\alpha_{n-12}$ again.

The problem reduces to $2 \leq n \leq 15$. In fact $n \in\{2,5,6,9,10,13,14\}$ by $n \equiv 1,2(\bmod 4)$. The cases $n=2,6,10,14$ reduce to $n=1,5,9,13$ respectively because the last even term of $\alpha_{n}$ can be ignored. For $n=5$ apply $\{4,5\} \mapsto 3$, then $\{3,3\} \mapsto 2$, then ignore the 2 occurrences of 2 . For $n=9$ ignore 6 first, then apply $\{5,7\} \mapsto 4,\{4,8\} \mapsto 4,\{3,9\} \mapsto 4$. Now ignore the 3 occurrences of 4 , then ignore 2 . Finally $n=13$ reduces to $n=10$ by $\{11,13\} \mapsto 8$ and ignoring 8 and 12. The proof is complete.

N8. Prove that for every prime $p>100$ and every integer $r$ there exist two integers $a$ and $b$ such that $p$ divides $a^{2}+b^{5}-r$.

Solution 1. Throughout the solution, all congruence relations are meant modulo $p$.
Fix $p$, and let $\mathcal{P}=\{0,1, \ldots, p-1\}$ be the set of residue classes modulo $p$. For every $r \in \mathcal{P}$, let $S_{r}=\left\{(a, b) \in \mathcal{P} \times \mathcal{P}: a^{2}+b^{5} \equiv r\right\}$, and let $s_{r}=\left|S_{r}\right|$. Our aim is to prove $s_{r}>0$ for all $r \in \mathcal{P}$.

We will use the well-known fact that for every residue class $r \in \mathcal{P}$ and every positive integer $k$, there are at most $k$ values $x \in \mathcal{P}$ such that $x^{k} \equiv r$.
Lemma. Let $N$ be the number of quadruples $(a, b, c, d) \in \mathcal{P}^{4}$ for which $a^{2}+b^{5} \equiv c^{2}+d^{5}$. Then

$$
\begin{equation*}
N=\sum_{r \in \mathcal{P}} s_{r}^{2} \tag{a}
\end{equation*}
$$

and

$$
\begin{equation*}
N \leq p\left(p^{2}+4 p-4\right) \tag{b}
\end{equation*}
$$

Proof. (a) For each residue class $r$ there exist exactly $s_{r}$ pairs $(a, b)$ with $a^{2}+b^{5} \equiv r$ and $s_{r}$ pairs $(c, d)$ with $c^{2}+d^{5} \equiv r$. So there are $s_{r}^{2}$ quadruples with $a^{2}+b^{5} \equiv c^{2}+d^{5} \equiv r$. Taking the sum over all $r \in \mathcal{P}$, the statement follows.
(b) Choose an arbitrary pair $(b, d) \in \mathcal{P}$ and look for the possible values of $a, c$.

1. Suppose that $b^{5} \equiv d^{5}$, and let $k$ be the number of such pairs $(b, d)$. The value $b$ can be chosen in $p$ different ways. For $b \equiv 0$ only $d=0$ has this property; for the nonzero values of $b$ there are at most 5 possible values for $d$. So we have $k \leq 1+5(p-1)=5 p-4$.

The values $a$ and $c$ must satisfy $a^{2} \equiv c^{2}$, so $a \equiv \pm c$, and there are exactly $2 p-1$ such pairs ( $a, c$ ).
2. Now suppose $b^{5} \not \equiv d^{5}$. In this case $a$ and $c$ must be distinct. By $(a-c)(a+c)=d^{5}-b^{5}$, the value of $a-c$ uniquely determines $a+c$ and thus $a$ and $c$ as well. Hence, there are $p-1$ suitable pairs $(a, c)$.

Thus, for each of the $k$ pairs $(b, d)$ with $b^{5} \equiv d^{5}$ there are $2 p-1$ pairs $(a, c)$, and for each of the other $p^{2}-k$ pairs $(b, d)$ there are $p-1$ pairs $(a, c)$. Hence,

$$
N=k(2 p-1)+\left(p^{2}-k\right)(p-1)=p^{2}(p-1)+k p \leq p^{2}(p-1)+(5 p-4) p=p\left(p^{2}+4 p-4\right)
$$

To prove the statement of the problem, suppose that $S_{r}=\emptyset$ for some $r \in \mathcal{P}$; obviously $r \not \equiv 0$. Let $T=\left\{x^{10}: x \in \mathcal{P} \backslash\{0\}\right\}$ be the set of nonzero 10 th powers modulo $p$. Since each residue class is the 10 th power of at most 10 elements in $\mathcal{P}$, we have $|T| \geq \frac{p-1}{10} \geq 4$ by $p>100$.

For every $t \in T$, we have $S_{t r}=\emptyset$. Indeed, if $(x, y) \in S_{t r}$ and $t \equiv z^{10}$ then

$$
\left(z^{-5} x\right)^{2}+\left(z^{-2} y\right)^{5} \equiv t^{-1}\left(x^{2}+y^{5}\right) \equiv r,
$$

so $\left(z^{-5} x, z^{-2} y\right) \in S_{r}$. So, there are at least $\frac{p-1}{10} \geq 4$ empty sets among $S_{1}, \ldots, S_{p-1}$, and there are at most $p-4$ nonzero values among $s_{0}, s_{2}, \ldots, s_{p-1}$. Then by the AM-QM inequality we obtain

$$
N=\sum_{r \in \mathcal{P} \backslash r T} s_{r}^{2} \geq \frac{1}{p-4}\left(\sum_{r \in \mathcal{P} \backslash r T} s_{r}\right)^{2}=\frac{|\mathcal{P} \times \mathcal{P}|^{2}}{p-4}=\frac{p^{4}}{p-4}>p\left(p^{2}+4 p-4\right)
$$

which is impossible by the lemma.

Solution 2. If $5 \nmid p-1$, then all modulo $p$ residue classes are complete fifth powers and the statement is trivial. So assume that $p=10 k+1$ where $k \geq 10$. Let $g$ be a primitive root modulo $p$.

We will use the following facts:
(F1) If some residue class $x$ is not quadratic then $x^{(p-1) / 2} \equiv-1(\bmod p)$.
(F2) For every integer $d$, as a simple corollary of the summation formula for geometric progressions,

$$
\sum_{i=0}^{2 k-1} g^{5 d i} \equiv\left\{\begin{array}{ll}
2 k & \text { if } 2 k \mid d \\
0 & \text { if } 2 k \nmid d
\end{array} \quad(\bmod p)\right.
$$

Suppose that, contrary to the statement, some modulo $p$ residue class $r$ cannot be expressed as $a^{2}+b^{5}$. Of course $r \not \equiv 0(\bmod p)$. By (F1) we have $\left(r-b^{5}\right)^{(p-1) / 2}=\left(r-b^{5}\right)^{5 k} \equiv-1(\bmod p)$ for all residue classes $b$.

For $t=1,2 \ldots, k-1$ consider the sums

$$
S(t)=\sum_{i=0}^{2 k-1}\left(r-g^{5 i}\right)^{5 k} g^{5 t i}
$$

By the indirect assumption and (F2),

$$
S(t)=\sum_{i=0}^{2 k-1}\left(r-\left(g^{i}\right)^{5}\right)^{5 k} g^{5 t i} \equiv \sum_{i=0}^{2 k-1}(-1) g^{5 t i} \equiv-\sum_{i=0}^{2 k-1} g^{5 t i} \equiv 0 \quad(\bmod p)
$$

because $2 k$ cannot divide $t$.
On the other hand, by the binomial theorem,

$$
\begin{aligned}
S(t) & =\sum_{i=0}^{2 k-1}\left(\sum_{j=0}^{5 k}\binom{5 k}{j} r^{5 k-j}\left(-g^{5 i}\right)^{j}\right) g^{5 t i}=\sum_{j=0}^{5 k}(-1)^{j}\binom{5 k}{j} r^{5 k-j}\left(\sum_{i=0}^{2 k-1} g^{5(j+t) i}\right) \equiv \\
& \equiv \sum_{j=0}^{5 k}(-1)^{j}\binom{5 k}{j} r^{5 k-j}\left\{\begin{array}{ll}
2 k & \text { if } 2 k \mid j+t \\
0 & \text { if } 2 k \nmid j+t
\end{array} \quad(\bmod p) .\right.
\end{aligned}
$$

Since $1 \leq j+t<6 k$, the number $2 k$ divides $j+t$ only for $j=2 k-t$ and $j=4 k-t$. Hence,

$$
\begin{gathered}
0 \equiv S(t) \equiv(-1)^{t}\left(\binom{5 k}{2 k-t} r^{3 k+t}+\binom{5 k}{4 k-t} r^{k+t}\right) \cdot 2 k \quad(\bmod p) \\
\binom{5 k}{2 k-t} r^{2 k}+\binom{5 k}{4 k-t} \equiv 0 \quad(\bmod p)
\end{gathered}
$$

Taking this for $t=1,2$ and eliminating $r$, we get

$$
\begin{aligned}
0 & \equiv\binom{5 k}{2 k-2}\left(\binom{5 k}{2 k-1} r^{2 k}+\binom{5 k}{4 k-1}\right)-\binom{5 k}{2 k-1}\left(\binom{5 k}{2 k-2} r^{2 k}+\binom{5 k}{4 k-2}\right) \\
& =\binom{5 k}{2 k-2}\binom{5 k}{4 k-1}-\binom{5 k}{2 k-1}\binom{5 k}{4 k-2} \\
& =\frac{(5 k)!^{2}}{(2 k-1)!(3 k+2)!(4 k-1)!(k+2)!}((2 k-1)(k+2)-(3 k+2)(4 k-1)) \\
& =\frac{-(5 k)!^{2} \cdot 2 k(5 k+1)}{(2 k-1)!(3 k+2)!(4 k-1)!(k+2)!}(\bmod p) .
\end{aligned}
$$

But in the last expression none of the numbers is divisible by $p=10 k+1$, a contradiction.

Comment 1. The argument in the second solution is valid whenever $k \geq 3$, that is for all primes $p=10 k+1$ except $p=11$. This is an exceptional case when the statement is not true; $r=7$ cannot be expressed as desired.

Comment 2. The statement is true in a more general setting: for every positive integer $n$, for all sufficiently large $p$, each residue class modulo $p$ can be expressed as $a^{2}+b^{n}$. Choosing $t=3$ would allow using the Cauchy-Davenport theorem (together with some analysis on the case of equality).

In the literature more general results are known. For instance, the statement easily follows from the Hasse-Weil bound.

# Shortlisted Problems with Solutions 

$54^{\text {th }}$ International Mathematical Olympiad Santa Marta, Colombia 2013

## Note of Confidentiality

## The Shortlisted Problems should be kept strictly confidential until IMO 2014.

## Contributing Countries

The Organizing Committee and the Problem Selection Committee of IMO 2013 thank the following 50 countries for contributing 149 problem proposals.

Argentina, Armenia, Australia, Austria, Belgium, Belarus, Brazil, Bulgaria, Croatia, Cyprus, Czech Republic, Denmark, El Salvador, Estonia, Finland, France, Georgia, Germany, Greece, Hungary, India, Indonesia, Iran, Ireland, Israel, Italy, Japan, Latvia, Lithuania, Luxembourg, Malaysia, Mexico, Netherlands, Nicaragua, Pakistan, Panama, Poland, Romania, Russia, Saudi Arabia, Serbia, Slovenia, Sweden, Switzerland, Tajikistan, Thailand, Turkey, U.S.A., Ukraine, United Kingdom

## Problem Selection Committee

Federico Ardila (chairman)
Ilya I. Bogdanov
Géza Kós
Carlos Gustavo Tamm de Araújo Moreira (Gugu)
Christian Reiher

## Problems

## Algebra

A1. Let $n$ be a positive integer and let $a_{1}, \ldots, a_{n-1}$ be arbitrary real numbers. Define the sequences $u_{0}, \ldots, u_{n}$ and $v_{0}, \ldots, v_{n}$ inductively by $u_{0}=u_{1}=v_{0}=v_{1}=1$, and

$$
u_{k+1}=u_{k}+a_{k} u_{k-1}, \quad v_{k+1}=v_{k}+a_{n-k} v_{k-1} \quad \text { for } k=1, \ldots, n-1 .
$$

Prove that $u_{n}=v_{n}$.
(France)
A2. Prove that in any set of 2000 distinct real numbers there exist two pairs $a>b$ and $c>d$ with $a \neq c$ or $b \neq d$, such that

$$
\left|\frac{a-b}{c-d}-1\right|<\frac{1}{100000} .
$$

(Lithuania)
A3. Let $\mathbb{Q}_{>0}$ be the set of positive rational numbers. Let $f: \mathbb{Q}_{>0} \rightarrow \mathbb{R}$ be a function satisfying the conditions

$$
f(x) f(y) \geqslant f(x y) \quad \text { and } \quad f(x+y) \geqslant f(x)+f(y)
$$

for all $x, y \in \mathbb{Q}_{>0}$. Given that $f(a)=a$ for some rational $a>1$, prove that $f(x)=x$ for all $x \in \mathbb{Q}_{>0}$.
(Bulgaria)
A4. Let $n$ be a positive integer, and consider a sequence $a_{1}, a_{2}, \ldots, a_{n}$ of positive integers. Extend it periodically to an infinite sequence $a_{1}, a_{2}, \ldots$ by defining $a_{n+i}=a_{i}$ for all $i \geqslant 1$. If

$$
a_{1} \leqslant a_{2} \leqslant \cdots \leqslant a_{n} \leqslant a_{1}+n
$$

and

$$
a_{a_{i}} \leqslant n+i-1 \quad \text { for } i=1,2, \ldots, n
$$

prove that

$$
a_{1}+\cdots+a_{n} \leqslant n^{2} .
$$

(Germany)
A5. Let $\mathbb{Z}_{\geqslant 0}$ be the set of all nonnegative integers. Find all the functions $f: \mathbb{Z}_{\geqslant 0} \rightarrow \mathbb{Z}_{\geqslant 0}$ satisfying the relation

$$
f(f(f(n)))=f(n+1)+1
$$

for all $n \in \mathbb{Z}_{\geqslant 0}$.
(Serbia)
A6. Let $m \neq 0$ be an integer. Find all polynomials $P(x)$ with real coefficients such that

$$
\left(x^{3}-m x^{2}+1\right) P(x+1)+\left(x^{3}+m x^{2}+1\right) P(x-1)=2\left(x^{3}-m x+1\right) P(x)
$$

for all real numbers $x$.

## Combinatorics

C1. Let $n$ be a positive integer. Find the smallest integer $k$ with the following property: Given any real numbers $a_{1}, \ldots, a_{d}$ such that $a_{1}+a_{2}+\cdots+a_{d}=n$ and $0 \leqslant a_{i} \leqslant 1$ for $i=1,2, \ldots, d$, it is possible to partition these numbers into $k$ groups (some of which may be empty) such that the sum of the numbers in each group is at most 1.
(Poland)
C2. In the plane, 2013 red points and 2014 blue points are marked so that no three of the marked points are collinear. One needs to draw $k$ lines not passing through the marked points and dividing the plane into several regions. The goal is to do it in such a way that no region contains points of both colors.

Find the minimal value of $k$ such that the goal is attainable for every possible configuration of 4027 points.
(Australia)
C3. A crazy physicist discovered a new kind of particle which he called an imon, after some of them mysteriously appeared in his lab. Some pairs of imons in the lab can be entangled, and each imon can participate in many entanglement relations. The physicist has found a way to perform the following two kinds of operations with these particles, one operation at a time.
( $i$ If some imon is entangled with an odd number of other imons in the lab, then the physicist can destroy it.
(ii) At any moment, he may double the whole family of imons in his lab by creating a copy $I^{\prime}$ of each imon $I$. During this procedure, the two copies $I^{\prime}$ and $J^{\prime}$ become entangled if and only if the original imons $I$ and $J$ are entangled, and each copy $I^{\prime}$ becomes entangled with its original imon $I$; no other entanglements occur or disappear at this moment.

Prove that the physicist may apply a sequence of such operations resulting in a family of imons, no two of which are entangled.
(Japan)
C4. Let $n$ be a positive integer, and let $A$ be a subset of $\{1, \ldots, n\}$. An $A$-partition of $n$ into $k$ parts is a representation of $n$ as a sum $n=a_{1}+\cdots+a_{k}$, where the parts $a_{1}, \ldots, a_{k}$ belong to $A$ and are not necessarily distinct. The number of different parts in such a partition is the number of (distinct) elements in the set $\left\{a_{1}, a_{2}, \ldots, a_{k}\right\}$.

We say that an $A$-partition of $n$ into $k$ parts is optimal if there is no $A$-partition of $n$ into $r$ parts with $r<k$. Prove that any optimal $A$-partition of $n$ contains at most $\sqrt[3]{6 n}$ different parts.
(Germany)
C5. Let $r$ be a positive integer, and let $a_{0}, a_{1}, \ldots$ be an infinite sequence of real numbers. Assume that for all nonnegative integers $m$ and $s$ there exists a positive integer $n \in[m+1, m+r]$ such that

$$
a_{m}+a_{m+1}+\cdots+a_{m+s}=a_{n}+a_{n+1}+\cdots+a_{n+s}
$$

Prove that the sequence is periodic, i. e. there exists some $p \geqslant 1$ such that $a_{n+p}=a_{n}$ for all $n \geqslant 0$.

C6. In some country several pairs of cities are connected by direct two-way flights. It is possible to go from any city to any other by a sequence of flights. The distance between two cities is defined to be the least possible number of flights required to go from one of them to the other. It is known that for any city there are at most 100 cities at distance exactly three from it. Prove that there is no city such that more than 2550 other cities have distance exactly four from it.
(Russia)
C7. Let $n \geqslant 2$ be an integer. Consider all circular arrangements of the numbers $0,1, \ldots, n$; the $n+1$ rotations of an arrangement are considered to be equal. A circular arrangement is called beautiful if, for any four distinct numbers $0 \leqslant a, b, c, d \leqslant n$ with $a+c=b+d$, the chord joining numbers $a$ and $c$ does not intersect the chord joining numbers $b$ and $d$.

Let $M$ be the number of beautiful arrangements of $0,1, \ldots, n$. Let $N$ be the number of pairs $(x, y)$ of positive integers such that $x+y \leqslant n$ and $\operatorname{gcd}(x, y)=1$. Prove that

$$
M=N+1
$$

(Russia)
C8. Players $A$ and $B$ play a paintful game on the real line. Player $A$ has a pot of paint with four units of black ink. A quantity $p$ of this ink suffices to blacken a (closed) real interval of length $p$. In every round, player $A$ picks some positive integer $m$ and provides $1 / 2^{m}$ units of ink from the pot. Player $B$ then picks an integer $k$ and blackens the interval from $k / 2^{m}$ to $(k+1) / 2^{m}$ (some parts of this interval may have been blackened before). The goal of player $A$ is to reach a situation where the pot is empty and the interval $[0,1]$ is not completely blackened.

Decide whether there exists a strategy for player $A$ to win in a finite number of moves.

## Geometry

G1. Let $A B C$ be an acute-angled triangle with orthocenter $H$, and let $W$ be a point on side $B C$. Denote by $M$ and $N$ the feet of the altitudes from $B$ and $C$, respectively. Denote by $\omega_{1}$ the circumcircle of $B W N$, and let $X$ be the point on $\omega_{1}$ which is diametrically opposite to $W$. Analogously, denote by $\omega_{2}$ the circumcircle of $C W M$, and let $Y$ be the point on $\omega_{2}$ which is diametrically opposite to $W$. Prove that $X, Y$ and $H$ are collinear.
(Thaliand)
G2. Let $\omega$ be the circumcircle of a triangle $A B C$. Denote by $M$ and $N$ the midpoints of the sides $A B$ and $A C$, respectively, and denote by $T$ the midpoint of the arc $B C$ of $\omega$ not containing $A$. The circumcircles of the triangles $A M T$ and $A N T$ intersect the perpendicular bisectors of $A C$ and $A B$ at points $X$ and $Y$, respectively; assume that $X$ and $Y$ lie inside the triangle $A B C$. The lines $M N$ and $X Y$ intersect at $K$. Prove that $K A=K T$.
(Iran)
G3. In a triangle $A B C$, let $D$ and $E$ be the feet of the angle bisectors of angles $A$ and $B$, respectively. A rhombus is inscribed into the quadrilateral $A E D B$ (all vertices of the rhombus lie on different sides of $A E D B$ ). Let $\varphi$ be the non-obtuse angle of the rhombus. Prove that $\varphi \leqslant \max \{\angle B A C, \angle A B C\}$.
(Serbia)
G4. Let $A B C$ be a triangle with $\angle B>\angle C$. Let $P$ and $Q$ be two different points on line $A C$ such that $\angle P B A=\angle Q B A=\angle A C B$ and $A$ is located between $P$ and $C$. Suppose that there exists an interior point $D$ of segment $B Q$ for which $P D=P B$. Let the ray $A D$ intersect the circle $A B C$ at $R \neq A$. Prove that $Q B=Q R$.
(Georgia)
G5. Let $A B C D E F$ be a convex hexagon with $A B=D E, B C=E F, C D=F A$, and $\angle A-\angle D=\angle C-\angle F=\angle E-\angle B$. Prove that the diagonals $A D, B E$, and $C F$ are concurrent.
(Ukraine)
G6. Let the excircle of the triangle $A B C$ lying opposite to $A$ touch its side $B C$ at the point $A_{1}$. Define the points $B_{1}$ and $C_{1}$ analogously. Suppose that the circumcentre of the triangle $A_{1} B_{1} C_{1}$ lies on the circumcircle of the triangle $A B C$. Prove that the triangle $A B C$ is right-angled.
(Russia)

## Number Theory

N1. Let $\mathbb{Z}_{>0}$ be the set of positive integers. Find all functions $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ such that

$$
m^{2}+f(n) \mid m f(m)+n
$$

for all positive integers $m$ and $n$.
(Malaysia)
N2. Prove that for any pair of positive integers $k$ and $n$ there exist $k$ positive integers $m_{1}, m_{2}, \ldots, m_{k}$ such that

$$
1+\frac{2^{k}-1}{n}=\left(1+\frac{1}{m_{1}}\right)\left(1+\frac{1}{m_{2}}\right) \cdots\left(1+\frac{1}{m_{k}}\right) .
$$

(Japan)
N3. Prove that there exist infinitely many positive integers $n$ such that the largest prime divisor of $n^{4}+n^{2}+1$ is equal to the largest prime divisor of $(n+1)^{4}+(n+1)^{2}+1$.
(Belgium)
N4. Determine whether there exists an infinite sequence of nonzero digits $a_{1}, a_{2}, a_{3}, \ldots$ and a positive integer $N$ such that for every integer $k>N$, the number $\overline{a_{k} a_{k-1} \ldots a_{1}}$ is a perfect square.
(Iran)
N5. Fix an integer $k \geqslant 2$. Two players, called Ana and Banana, play the following game of numbers: Initially, some integer $n \geqslant k$ gets written on the blackboard. Then they take moves in turn, with Ana beginning. A player making a move erases the number $m$ just written on the blackboard and replaces it by some number $m^{\prime}$ with $k \leqslant m^{\prime}<m$ that is coprime to $m$. The first player who cannot move anymore loses.

An integer $n \geqslant k$ is called good if Banana has a winning strategy when the initial number is $n$, and bad otherwise.

Consider two integers $n, n^{\prime} \geqslant k$ with the property that each prime number $p \leqslant k$ divides $n$ if and only if it divides $n^{\prime}$. Prove that either both $n$ and $n^{\prime}$ are good or both are bad.
(Italy)
N6. Determine all functions $f: \mathbb{Q} \longrightarrow \mathbb{Z}$ satisfying

$$
f\left(\frac{f(x)+a}{b}\right)=f\left(\frac{x+a}{b}\right)
$$

for all $x \in \mathbb{Q}, a \in \mathbb{Z}$, and $b \in \mathbb{Z}_{>0}$. (Here, $\mathbb{Z}_{>0}$ denotes the set of positive integers.)
(Israel)
N7. Let $\nu$ be an irrational positive number, and let $m$ be a positive integer. A pair $(a, b)$ of positive integers is called good if

$$
a\lceil b \nu\rceil-b\lfloor a \nu\rfloor=m .
$$

A good pair $(a, b)$ is called excellent if neither of the pairs $(a-b, b)$ and $(a, b-a)$ is good. (As usual, by $\lfloor x\rfloor$ and $\lceil x\rceil$ we denote the integer numbers such that $x-1<\lfloor x\rfloor \leqslant x$ and $x \leqslant\lceil x\rceil<x+1$.)

Prove that the number of excellent pairs is equal to the sum of the positive divisors of $m$.
(U.S.A.)

## Solutions

## Algebra

A1. Let $n$ be a positive integer and let $a_{1}, \ldots, a_{n-1}$ be arbitrary real numbers. Define the sequences $u_{0}, \ldots, u_{n}$ and $v_{0}, \ldots, v_{n}$ inductively by $u_{0}=u_{1}=v_{0}=v_{1}=1$, and

$$
u_{k+1}=u_{k}+a_{k} u_{k-1}, \quad v_{k+1}=v_{k}+a_{n-k} v_{k-1} \quad \text { for } k=1, \ldots, n-1
$$

Prove that $u_{n}=v_{n}$.
(France)
Solution 1. We prove by induction on $k$ that

$$
\begin{equation*}
u_{k}=\sum_{\substack{0<i_{1}<\ldots<i_{t}<k, i_{j+1}-i_{j} \geqslant 2}} a_{i_{1}} \ldots a_{i_{t}} . \tag{1}
\end{equation*}
$$

Note that we have one trivial summand equal to 1 (which corresponds to $t=0$ and the empty sequence, whose product is 1 ).

For $k=0,1$ the sum on the right-hand side only contains the empty product, so (1) holds due to $u_{0}=u_{1}=1$. For $k \geqslant 1$, assuming the result is true for $0,1, \ldots, k$, we have

$$
\begin{aligned}
u_{k+1} & =\sum_{\substack{0<i_{1}<\ldots<i_{t}<k, i_{j+1}-i_{j} \geqslant 2}} a_{i_{1}} \ldots a_{i_{t}}+\sum_{\substack{0<i_{1}<\ldots<i_{t}<k-1, i_{j+1}-i_{j} \geqslant 2}} a_{i_{1}} \ldots a_{i_{t}} \cdot a_{k} \\
& =\sum_{\substack{\left.0<i_{1}<\ldots<i_{t}<k+1, i_{j+1}+i_{j} \geqslant 2, k \notin i_{1}, \ldots, i_{t}\right\}}} \ldots a_{i_{t}}+\sum_{\substack{0<i_{1}<\ldots<i_{t}<k+1, i_{j+1}-1-i_{j} \geqslant 2, k \in\left\{i_{1}, \ldots, i_{t}\right\}}} a_{i_{1}} \ldots a_{i_{t}} \\
& =\sum_{\substack{0<i_{1}<\ldots<i_{t}<k+1, i_{j+1}-i_{j} \geqslant 2}} a_{i_{1}} \ldots a_{i_{t}},
\end{aligned}
$$

as required.
Applying (1) to the sequence $b_{1}, \ldots, b_{n}$ given by $b_{k}=a_{n-k}$ for $1 \leqslant k \leqslant n$, we get

$$
\begin{equation*}
v_{k}=\sum_{\substack{0<i_{1}<\ldots<i_{t}<k, i_{j+1}-i_{j} \geqslant 2}} b_{i_{1}} \ldots b_{i_{t}}=\sum_{\substack{n>i_{1}>\ldots>i_{t}>n-k, i_{j}-i_{j+1} \geqslant 2}} a_{i_{1}} \ldots a_{i_{t}} . \tag{2}
\end{equation*}
$$

For $k=n$ the expressions (1) and (2) coincide, so indeed $u_{n}=v_{n}$.
Solution 2. Define recursively a sequence of multivariate polynomials by

$$
P_{0}=P_{1}=1, \quad P_{k+1}\left(x_{1}, \ldots, x_{k}\right)=P_{k}\left(x_{1}, \ldots, x_{k-1}\right)+x_{k} P_{k-1}\left(x_{1}, \ldots, x_{k-2}\right),
$$

so $P_{n}$ is a polynomial in $n-1$ variables for each $n \geqslant 1$. Two easy inductive arguments show that

$$
u_{n}=P_{n}\left(a_{1}, \ldots, a_{n-1}\right), \quad v_{n}=P_{n}\left(a_{n-1}, \ldots, a_{1}\right),
$$

so we need to prove $P_{n}\left(x_{1}, \ldots, x_{n-1}\right)=P_{n}\left(x_{n-1}, \ldots, x_{1}\right)$ for every positive integer $n$. The cases $n=1,2$ are trivial, and the cases $n=3,4$ follow from $P_{3}(x, y)=1+x+y$ and $P_{4}(x, y, z)=$ $1+x+y+z+x z$.

Now we proceed by induction, assuming that $n \geqslant 5$ and the claim hold for all smaller cases. Using $F(a, b)$ as an abbreviation for $P_{|a-b|+1}\left(x_{a}, \ldots, x_{b}\right)$ (where the indices $a, \ldots, b$ can be either in increasing or decreasing order),

$$
\begin{aligned}
F(n, 1) & =F(n, 2)+x_{1} F(n, 3)=F(2, n)+x_{1} F(3, n) \\
& =\left(F(2, n-1)+x_{n} F(2, n-2)\right)+x_{1}\left(F(3, n-1)+x_{n} F(3, n-2)\right) \\
& =\left(F(n-1,2)+x_{1} F(n-1,3)\right)+x_{n}\left(F(n-2,2)+x_{1} F(n-2,3)\right) \\
& =F(n-1,1)+x_{n} F(n-2,1)=F(1, n-1)+x_{n} F(1, n-2) \\
& =F(1, n),
\end{aligned}
$$

as we wished to show.
Solution 3. Using matrix notation, we can rewrite the recurrence relation as

$$
\binom{u_{k+1}}{u_{k+1}-u_{k}}=\binom{u_{k}+a_{k} u_{k-1}}{a_{k} u_{k-1}}=\left(\begin{array}{cc}
1+a_{k} & -a_{k} \\
a_{k} & -a_{k}
\end{array}\right)\binom{u_{k}}{u_{k}-u_{k-1}}
$$

for $1 \leqslant k \leqslant n-1$, and similarly

$$
\left(v_{k+1} ; v_{k}-v_{k+1}\right)=\left(v_{k}+a_{n-k} v_{k-1} ;-a_{n-k} v_{k-1}\right)=\left(v_{k} ; v_{k-1}-v_{k}\right)\left(\begin{array}{cc}
1+a_{n-k} & -a_{n-k} \\
a_{n-k} & -a_{n-k}
\end{array}\right)
$$

for $1 \leqslant k \leqslant n-1$. Hence, introducing the $2 \times 2$ matrices $A_{k}=\left(\begin{array}{cc}1+a_{k} & -a_{k} \\ a_{k} & -a_{k}\end{array}\right)$ we have

$$
\binom{u_{k+1}}{u_{k+1}-u_{k}}=A_{k}\binom{u_{k}}{u_{k}-u_{k-1}} \quad \text { and } \quad\left(v_{k+1} ; v_{k}-v_{k+1}\right)=\left(v_{k} ; v_{k-1}-v_{k}\right) A_{n-k} .
$$

for $1 \leqslant k \leqslant n-1$. Since $\binom{u_{1}}{u_{1}-u_{0}}=\binom{1}{0}$ and $\left(v_{1} ; v_{0}-v_{1}\right)=(1 ; 0)$, we get

$$
\binom{u_{n}}{u_{n}-u_{n-1}}=A_{n-1} A_{n-2} \cdots A_{1} \cdot\binom{1}{0} \quad \text { and } \quad\left(v_{n} ; v_{n-1}-v_{n}\right)=(1 ; 0) \cdot A_{n-1} A_{n-2} \cdots A_{1} .
$$

It follows that

$$
\left(u_{n}\right)=(1 ; 0)\binom{u_{n}}{u_{n}-u_{n-1}}=(1 ; 0) \cdot A_{n-1} A_{n-2} \cdots A_{1} \cdot\binom{1}{0}=\left(v_{n} ; v_{n-1}-v_{n}\right)\binom{1}{0}=\left(v_{n}\right) .
$$

Comment 1. These sequences are related to the Fibonacci sequence; when $a_{1}=\cdots=a_{n-1}=1$, we have $u_{k}=v_{k}=F_{k+1}$, the $(k+1)$ st Fibonacci number. Also, for every positive integer $k$, the polynomial $P_{k}\left(x_{1}, \ldots, x_{k-1}\right)$ from Solution 2 is the sum of $F_{k+1}$ monomials.

Comment 2. One may notice that the condition is equivalent to

$$
\frac{u_{k+1}}{u_{k}}=1+\frac{a_{k}}{1+\frac{a_{k-1}}{1+\ldots+\frac{a_{2}}{1+a_{1}}}} \quad \text { and } \quad \frac{v_{k+1}}{v_{k}}=1+\frac{a_{n-k}}{1+\frac{a_{n-k+1}}{1+\ldots+\frac{a_{n-2}}{1+a_{n-1}}}}
$$

so the problem claims that the corresponding continued fractions for $u_{n} / u_{n-1}$ and $v_{n} / v_{n-1}$ have the same numerator.

Comment 3. An alternative variant of the problem is the following.
Let $n$ be a positive integer and let $a_{1}, \ldots, a_{n-1}$ be arbitrary real numbers. Define the sequences $u_{0}, \ldots, u_{n}$ and $v_{0}, \ldots, v_{n}$ inductively by $u_{0}=v_{0}=0, u_{1}=v_{1}=1$, and

$$
u_{k+1}=a_{k} u_{k}+u_{k-1}, \quad v_{k+1}=a_{n-k} v_{k}+v_{k-1} \quad \text { for } k=1, \ldots, n-1 .
$$

Prove that $u_{n}=v_{n}$.
All three solutions above can be reformulated to prove this statement; one may prove

$$
u_{n}=v_{n}=\sum_{\substack{0=i_{0}<i_{1}<\ldots<i_{t}=n, i_{j+1}-i_{j} \text { is odd }}} a_{i_{1}} \ldots a_{i_{t-1}} \quad \text { for } n>0
$$

or observe that

$$
\binom{u_{k+1}}{u_{k}}=\left(\begin{array}{cc}
a_{k} & 1 \\
1 & 0
\end{array}\right)\binom{u_{k}}{u_{k-1}} \quad \text { and } \quad\left(v_{k+1} ; v_{k}\right)=\left(v_{k} ; v_{k-1}\right)\left(\begin{array}{cc}
a_{k} & 1 \\
1 & 0
\end{array}\right) .
$$

Here we have

$$
\frac{u_{k+1}}{u_{k}}=a_{k}+\frac{1}{a_{k-1}+\frac{1}{a_{k-2}+\ldots+\frac{1}{a_{1}}}}=\left[a_{k} ; a_{k-1}, \ldots, a_{1}\right]
$$

and

$$
\frac{v_{k+1}}{v_{k}}=a_{n-k}+\frac{1}{a_{n-k+1}+\frac{1}{a_{n-k+2}+\ldots+\frac{1}{a_{n-1}}}}=\left[a_{n-k} ; a_{n-k+1}, \ldots, a_{n-1}\right],
$$

so this alternative statement is equivalent to the known fact that the continued fractions $\left[a_{n-1} ; a_{n-2}, \ldots, a_{1}\right]$ and $\left[a_{1} ; a_{2}, \ldots, a_{n-1}\right]$ have the same numerator.

A2. Prove that in any set of 2000 distinct real numbers there exist two pairs $a>b$ and $c>d$ with $a \neq c$ or $b \neq d$, such that

$$
\left|\frac{a-b}{c-d}-1\right|<\frac{1}{100000}
$$

(Lithuania)
Solution. For any set $S$ of $n=2000$ distinct real numbers, let $D_{1} \leqslant D_{2} \leqslant \cdots \leqslant D_{m}$ be the distances between them, displayed with their multiplicities. Here $m=n(n-1) / 2$. By rescaling the numbers, we may assume that the smallest distance $D_{1}$ between two elements of $S$ is $D_{1}=1$. Let $D_{1}=1=y-x$ for $x, y \in S$. Evidently $D_{m}=v-u$ is the difference between the largest element $v$ and the smallest element $u$ of $S$.

If $D_{i+1} / D_{i}<1+10^{-5}$ for some $i=1,2, \ldots, m-1$ then the required inequality holds, because $0 \leqslant D_{i+1} / D_{i}-1<10^{-5}$. Otherwise, the reverse inequality

$$
\frac{D_{i+1}}{D_{i}} \geqslant 1+\frac{1}{10^{5}}
$$

holds for each $i=1,2, \ldots, m-1$, and therefore

$$
v-u=D_{m}=\frac{D_{m}}{D_{1}}=\frac{D_{m}}{D_{m-1}} \cdots \frac{D_{3}}{D_{2}} \cdot \frac{D_{2}}{D_{1}} \geqslant\left(1+\frac{1}{10^{5}}\right)^{m-1} .
$$

From $m-1=n(n-1) / 2-1=1000 \cdot 1999-1>19 \cdot 10^{5}$, together with the fact that for all $n \geqslant 1$, $\left(1+\frac{1}{n}\right)^{n} \geqslant 1+\binom{n}{1} \cdot \frac{1}{n}=2$, we get

$$
\left(1+\frac{1}{10^{5}}\right)^{19 \cdot 10^{5}}=\left(\left(1+\frac{1}{10^{5}}\right)^{10^{5}}\right)^{19} \geqslant 2^{19}=2^{9} \cdot 2^{10}>500 \cdot 1000>2 \cdot 10^{5}
$$

and so $v-u=D_{m}>2 \cdot 10^{5}$.
Since the distance of $x$ to at least one of the numbers $u, v$ is at least $(u-v) / 2>10^{5}$, we have

$$
|x-z|>10^{5}
$$

for some $z \in\{u, v\}$. Since $y-x=1$, we have either $z>y>x$ (if $z=v$ ) or $y>x>z$ (if $z=u$ ). If $z>y>x$, selecting $a=z, b=y, c=z$ and $d=x$ (so that $b \neq d$ ), we obtain

$$
\left|\frac{a-b}{c-d}-1\right|=\left|\frac{z-y}{z-x}-1\right|=\left|\frac{x-y}{z-x}\right|=\frac{1}{z-x}<10^{-5} .
$$

Otherwise, if $y>x>z$, we may choose $a=y, b=z, c=x$ and $d=z$ (so that $a \neq c$ ), and obtain

$$
\left|\frac{a-b}{c-d}-1\right|=\left|\frac{y-z}{x-z}-1\right|=\left|\frac{y-x}{x-z}\right|=\frac{1}{x-z}<10^{-5} .
$$

The desired result follows.

Comment. As the solution shows, the numbers 2000 and $\frac{1}{100000}$ appearing in the statement of the problem may be replaced by any $n \in \mathbb{Z}_{>0}$ and $\delta>0$ satisfying

$$
\delta(1+\delta)^{n(n-1) / 2-1}>2
$$

A3. Let $\mathbb{Q}_{>0}$ be the set of positive rational numbers. Let $f: \mathbb{Q}_{>0} \rightarrow \mathbb{R}$ be a function satisfying the conditions

$$
\begin{align*}
& f(x) f(y) \geqslant f(x y)  \tag{1}\\
& f(x+y) \geqslant f(x)+f(y) \tag{2}
\end{align*}
$$

for all $x, y \in \mathbb{Q}_{>0}$. Given that $f(a)=a$ for some rational $a>1$, prove that $f(x)=x$ for all $x \in \mathbb{Q}_{>0}$.
(Bulgaria)
Solution. Denote by $\mathbb{Z}_{>0}$ the set of positive integers.
Plugging $x=1, y=a$ into (1) we get $f(1) \geqslant 1$. Next, by an easy induction on $n$ we get from (2) that

$$
\begin{equation*}
f(n x) \geqslant n f(x) \quad \text { for all } n \in \mathbb{Z}_{>0} \text { and } x \in \mathbb{Q}_{>0} \tag{3}
\end{equation*}
$$

In particular, we have

$$
\begin{equation*}
f(n) \geqslant n f(1) \geqslant n \quad \text { for all } n \in \mathbb{Z}_{>0} \tag{4}
\end{equation*}
$$

From (1) again we have $f(m / n) f(n) \geqslant f(m)$, so $f(q)>0$ for all $q \in \mathbb{Q}_{>0}$.
Now, (2) implies that $f$ is strictly increasing; this fact together with (4) yields

$$
f(x) \geqslant f(\lfloor x\rfloor) \geqslant\lfloor x\rfloor>x-1 \quad \text { for all } x \geqslant 1
$$

By an easy induction we get from (1) that $f(x)^{n} \geqslant f\left(x^{n}\right)$, so

$$
f(x)^{n} \geqslant f\left(x^{n}\right)>x^{n}-1 \quad \Longrightarrow \quad f(x) \geqslant \sqrt[n]{x^{n}-1} \quad \text { for all } x>1 \text { and } n \in \mathbb{Z}_{>0}
$$

This yields

$$
\begin{equation*}
f(x) \geqslant x \quad \text { for every } x>1 \tag{5}
\end{equation*}
$$

(Indeed, if $x>y>1$ then $x^{n}-y^{n}=(x-y)\left(x^{n-1}+x^{n-2} y+\cdots+y^{n}\right)>n(x-y)$, so for a large $n$ we have $x^{n}-1>y^{n}$ and thus $f(x)>y$.)

Now, (1) and (5) give $a^{n}=f(a)^{n} \geqslant f\left(a^{n}\right) \geqslant a^{n}$, so $f\left(a^{n}\right)=a^{n}$. Now, for $x>1$ let us choose $n \in \mathbb{Z}_{>0}$ such that $a^{n}-x>1$. Then by (2) and (5) we get

$$
a^{n}=f\left(a^{n}\right) \geqslant f(x)+f\left(a^{n}-x\right) \geqslant x+\left(a^{n}-x\right)=a^{n}
$$

and therefore $f(x)=x$ for $x>1$. Finally, for every $x \in \mathbb{Q}_{>0}$ and every $n \in \mathbb{Z}_{>0}$, from (1) and (3) we get

$$
n f(x)=f(n) f(x) \geqslant f(n x) \geqslant n f(x)
$$

which gives $f(n x)=n f(x)$. Therefore $f(m / n)=f(m) / n=m / n$ for all $m, n \in \mathbb{Z}_{>0}$.
Comment. The condition $f(a)=a>1$ is essential. Indeed, for $b \geqslant 1$ the function $f(x)=b x^{2}$ satisfies (1) and (2) for all $x, y \in \mathbb{Q}_{>0}$, and it has a unique fixed point $1 / b \leqslant 1$.

A4. Let $n$ be a positive integer, and consider a sequence $a_{1}, a_{2}, \ldots, a_{n}$ of positive integers. Extend it periodically to an infinite sequence $a_{1}, a_{2}, \ldots$ by defining $a_{n+i}=a_{i}$ for all $i \geqslant 1$. If

$$
\begin{equation*}
a_{1} \leqslant a_{2} \leqslant \cdots \leqslant a_{n} \leqslant a_{1}+n \tag{1}
\end{equation*}
$$

and

$$
\begin{equation*}
a_{a_{i}} \leqslant n+i-1 \quad \text { for } i=1,2, \ldots, n \tag{2}
\end{equation*}
$$

prove that

$$
a_{1}+\cdots+a_{n} \leqslant n^{2} .
$$

(Germany)
Solution 1. First, we claim that

$$
\begin{equation*}
a_{i} \leqslant n+i-1 \quad \text { for } i=1,2, \ldots, n \text {. } \tag{3}
\end{equation*}
$$

Assume contrariwise that $i$ is the smallest counterexample. From $a_{n} \geqslant a_{n-1} \geqslant \cdots \geqslant a_{i} \geqslant n+i$ and $a_{a_{i}} \leqslant n+i-1$, taking into account the periodicity of our sequence, it follows that

$$
\begin{equation*}
a_{i} \text { cannot be congruent to } i, i+1, \ldots, n-1, \text { or } n(\bmod n) . \tag{4}
\end{equation*}
$$

Thus our assumption that $a_{i} \geqslant n+i$ implies the stronger statement that $a_{i} \geqslant 2 n+1$, which by $a_{1}+n \geqslant a_{n} \geqslant a_{i}$ gives $a_{1} \geqslant n+1$. The minimality of $i$ then yields $i=1$, and (4) becomes contradictory. This establishes our first claim.

In particular we now know that $a_{1} \leqslant n$. If $a_{n} \leqslant n$, then $a_{1} \leqslant \cdots \leqslant \cdots a_{n} \leqslant n$ and the desired inequality holds trivially. Otherwise, consider the number $t$ with $1 \leqslant t \leqslant n-1$ such that

$$
\begin{equation*}
a_{1} \leqslant a_{2} \leqslant \ldots \leqslant a_{t} \leqslant n<a_{t+1} \leqslant \ldots \leqslant a_{n} . \tag{5}
\end{equation*}
$$

Since $1 \leqslant a_{1} \leqslant n$ and $a_{a_{1}} \leqslant n$ by (2), we have $a_{1} \leqslant t$ and hence $a_{n} \leqslant n+t$. Therefore if for each positive integer $i$ we let $b_{i}$ be the number of indices $j \in\{t+1, \ldots, n\}$ satisfying $a_{j} \geqslant n+i$, we have

$$
b_{1} \geqslant b_{2} \geqslant \ldots \geqslant b_{t} \geqslant b_{t+1}=0 .
$$

Next we claim that $a_{i}+b_{i} \leqslant n$ for $1 \leqslant i \leqslant t$. Indeed, by $n+i-1 \geqslant a_{a_{i}}$ and $a_{i} \leqslant n$, each $j$ with $a_{j} \geqslant n+i$ (thus $a_{j}>a_{a_{i}}$ ) belongs to $\left\{a_{i}+1, \ldots, n\right\}$, and for this reason $b_{i} \leqslant n-a_{i}$.

It follows from the definition of the $b_{i} \mathrm{~S}$ and (5) that

$$
a_{t+1}+\ldots+a_{n} \leqslant n(n-t)+b_{1}+\ldots+b_{t} .
$$

Adding $a_{1}+\ldots+a_{t}$ to both sides and using that $a_{i}+b_{i} \leqslant n$ for $1 \leqslant i \leqslant t$, we get

$$
a_{1}+a_{2}+\cdots+a_{n} \leqslant n(n-t)+n t=n^{2}
$$

as we wished to prove.

Solution 2. In the first quadrant of an infinite grid, consider the increasing "staircase" obtained by shading in dark the bottom $a_{i}$ cells of the $i$ th column for $1 \leqslant i \leqslant n$. We will prove that there are at most $n^{2}$ dark cells.

To do it, consider the $n \times n$ square $S$ in the first quadrant with a vertex at the origin. Also consider the $n \times n$ square directly to the left of $S$. Starting from its lower left corner, shade in light the leftmost $a_{j}$ cells of the $j$ th row for $1 \leqslant j \leqslant n$. Equivalently, the light shading is obtained by reflecting the dark shading across the line $x=y$ and translating it $n$ units to the left. The figure below illustrates this construction for the sequence $6,6,6,7,7,7,8,12,12,14$.


We claim that there is no cell in $S$ which is both dark and light. Assume, contrariwise, that there is such a cell in column $i$. Consider the highest dark cell in column $i$ which is inside $S$. Since it is above a light cell and inside $S$, it must be light as well. There are two cases:

Case 1. $a_{i} \leqslant n$
If $a_{i} \leqslant n$ then this dark and light cell is $\left(i, a_{i}\right)$, as highlighted in the figure. However, this is the $(n+i)$-th cell in row $a_{i}$, and we only shaded $a_{a_{i}}<n+i$ light cells in that row, a contradiction.

Case 2. $a_{i} \geqslant n+1$
If $a_{i} \geqslant n+1$, this dark and light cell is $(i, n)$. This is the $(n+i)$-th cell in row $n$ and we shaded $a_{n} \leqslant a_{1}+n$ light cells in this row, so we must have $i \leqslant a_{1}$. But $a_{1} \leqslant a_{a_{1}} \leqslant n$ by (1) and (2), so $i \leqslant a_{1}$ implies $a_{i} \leqslant a_{a_{1}} \leqslant n$, contradicting our assumption.

We conclude that there are no cells in $S$ which are both dark and light. It follows that the number of shaded cells in $S$ is at most $n^{2}$.

Finally, observe that if we had a light cell to the right of $S$, then by symmetry we would have a dark cell above $S$, and then the cell $(n, n)$ would be dark and light. It follows that the number of light cells in $S$ equals the number of dark cells outside of $S$, and therefore the number of shaded cells in $S$ equals $a_{1}+\cdots+a_{n}$. The desired result follows.

Solution 3. As in Solution 1, we first establish that $a_{i} \leqslant n+i-1$ for $1 \leqslant i \leqslant n$. Now define $c_{i}=\max \left(a_{i}, i\right)$ for $1 \leqslant i \leqslant n$ and extend the sequence $c_{1}, c_{2}, \ldots$ periodically modulo $n$. We claim that this sequence also satisfies the conditions of the problem.

For $1 \leqslant i<j \leqslant n$ we have $a_{i} \leqslant a_{j}$ and $i<j$, so $c_{i} \leqslant c_{j}$. Also $a_{n} \leqslant a_{1}+n$ and $n<1+n$ imply $c_{n} \leqslant c_{1}+n$. Finally, the definitions imply that $c_{c_{i}} \in\left\{a_{a_{i}}, a_{i}, a_{i}-n, i\right\}$ so $c_{c_{i}} \leqslant n+i-1$ by (2) and (3). This establishes (1) and (2) for $c_{1}, c_{2}, \ldots$..

Our new sequence has the additional property that

$$
\begin{equation*}
c_{i} \geqslant i \quad \text { for } i=1,2, \ldots, n, \tag{6}
\end{equation*}
$$

which allows us to construct the following visualization: Consider $n$ equally spaced points on a circle, sequentially labelled $1,2, \ldots, n(\bmod n)$, so point $k$ is also labelled $n+k$. We draw arrows from vertex $i$ to vertices $i+1, \ldots, c_{i}$ for $1 \leqslant i \leqslant n$, keeping in mind that $c_{i} \geqslant i$ by (6). Since $c_{i} \leqslant n+i-1$ by (3), no arrow will be drawn twice, and there is no arrow from a vertex to itself. The total number of arrows is

$$
\text { number of arrows }=\sum_{i=1}^{n}\left(c_{i}-i\right)=\sum_{i=1}^{n} c_{i}-\binom{n+1}{2}
$$

Now we show that we never draw both arrows $i \rightarrow j$ and $j \rightarrow i$ for $1 \leqslant i<j \leqslant n$. Assume contrariwise. This means, respectively, that

$$
i<j \leqslant c_{i} \quad \text { and } \quad j<n+i \leqslant c_{j} .
$$

We have $n+i \leqslant c_{j} \leqslant c_{1}+n$ by (1), so $i \leqslant c_{1}$. Since $c_{1} \leqslant n$ by (3), this implies that $c_{i} \leqslant c_{c_{1}} \leqslant n$ using (1) and (3). But then, using (1) again, $j \leqslant c_{i} \leqslant n$ implies $c_{j} \leqslant c_{c_{i}}$, which combined with $n+i \leqslant c_{j}$ gives us that $n+i \leqslant c_{c_{i}}$. This contradicts (2).

This means that the number of arrows is at most $\binom{n}{2}$, which implies that

$$
\sum_{i=1}^{n} c_{i} \leqslant\binom{ n}{2}+\binom{n+1}{2}=n^{2}
$$

Recalling that $a_{i} \leqslant c_{i}$ for $1 \leqslant i \leqslant n$, the desired inequality follows.
Comment 1. We sketch an alternative proof by induction. Begin by verifying the initial case $n=1$ and the simple cases when $a_{1}=1, a_{1}=n$, or $a_{n} \leqslant n$. Then, as in Solution 1, consider the index $t$ such that $a_{1} \leqslant \cdots \leqslant a_{t} \leqslant n<a_{t+1} \leqslant \cdots \leqslant a_{n}$. Observe again that $a_{1} \leqslant t$. Define the sequence $d_{1}, \ldots, d_{n-1}$ by

$$
d_{i}= \begin{cases}a_{i+1}-1 & \text { if } i \leqslant t-1 \\ a_{i+1}-2 & \text { if } i \geqslant t\end{cases}
$$

and extend it periodically modulo $n-1$. One may verify that this sequence also satisfies the hypotheses of the problem. The induction hypothesis then gives $d_{1}+\cdots+d_{n-1} \leqslant(n-1)^{2}$, which implies that

$$
\sum_{i=1}^{n} a_{i}=a_{1}+\sum_{i=2}^{t}\left(d_{i-1}+1\right)+\sum_{i=t+1}^{n}\left(d_{i-1}+2\right) \leqslant t+(t-1)+2(n-t)+(n-1)^{2}=n^{2}
$$

Comment 2. One unusual feature of this problem is that there are many different sequences for which equality holds. The discovery of such optimal sequences is not difficult, and it is useful in guiding the steps of a proof.

In fact, Solution 2 gives a complete description of the optimal sequences. Start with any lattice path $P$ from the lower left to the upper right corner of the $n \times n$ square $S$ using only steps up and right, such that the total number of steps along the left and top edges of $S$ is at least $n$. Shade the cells of $S$ below $P$ dark, and the cells of $S$ above $P$ light. Now reflect the light shape across the line $x=y$ and shift it up $n$ units, and shade it dark. As Solution 2 shows, the dark region will then correspond to an optimal sequence, and every optimal sequence arises in this way.

A5. Let $\mathbb{Z}_{\geqslant 0}$ be the set of all nonnegative integers. Find all the functions $f: \mathbb{Z}_{\geqslant 0} \rightarrow \mathbb{Z}_{\geqslant 0}$ satisfying the relation

$$
\begin{equation*}
f(f(f(n)))=f(n+1)+1 \tag{*}
\end{equation*}
$$

for all $n \in \mathbb{Z}_{\geqslant 0}$.
(Serbia)
Answer. There are two such functions: $f(n)=n+1$ for all $n \in \mathbb{Z}_{\geqslant 0}$, and

$$
f(n)=\left\{\begin{array}{ll}
n+1, & n \equiv 0(\bmod 4) \text { or } n \equiv 2(\bmod 4),  \tag{1}\\
n+5, & n \equiv 1(\bmod 4), \\
n-3, & n \equiv 3(\bmod 4)
\end{array} \quad \text { for all } n \in \mathbb{Z}_{\geqslant 0}\right.
$$

Throughout all the solutions, we write $h^{k}(x)$ to abbreviate the $k$ th iteration of function $h$, so $h^{0}$ is the identity function, and $h^{k}(x)=\underbrace{h(\ldots h}_{k \text { times }}(x) \ldots))$ for $k \geqslant 1$.
Solution 1. To start, we get from (*) that

$$
f^{4}(n)=f\left(f^{3}(n)\right)=f(f(n+1)+1) \quad \text { and } \quad f^{4}(n+1)=f^{3}(f(n+1))=f(f(n+1)+1)+1
$$

thus

$$
\begin{equation*}
f^{4}(n)+1=f^{4}(n+1) . \tag{2}
\end{equation*}
$$

I. Let us denote by $R_{i}$ the range of $f^{i}$; note that $R_{0}=\mathbb{Z}_{\geqslant 0}$ since $f^{0}$ is the identity function. Obviously, $R_{0} \supseteq R_{1} \supseteq \ldots$ Next, from (2) we get that if $a \in R_{4}$ then also $a+1 \in R_{4}$. This implies that $\mathbb{Z}_{\geqslant 0} \backslash R_{4}$ - and hence $\mathbb{Z}_{\geqslant 0} \backslash R_{1}$ - is finite. In particular, $R_{1}$ is unbounded.

Assume that $f(m)=f(n)$ for some distinct $m$ and $n$. Then from (*) we obtain $f(m+1)=$ $f(n+1)$; by an easy induction we then get that $f(m+c)=f(n+c)$ for every $c \geqslant 0$. So the function $f(k)$ is periodic with period $|m-n|$ for $k \geqslant m$, and thus $R_{1}$ should be bounded, which is false. So, $f$ is injective.
II. Denote now $S_{i}=R_{i-1} \backslash R_{i}$; all these sets are finite for $i \leqslant 4$. On the other hand, by the injectivity we have $n \in S_{i} \Longleftrightarrow f(n) \in S_{i+1}$. By the injectivity again, $f$ implements a bijection between $S_{i}$ and $S_{i+1}$, thus $\left|S_{1}\right|=\left|S_{2}\right|=\ldots$; denote this common cardinality by $k$. If $0 \in R_{3}$ then $0=f(f(f(n)))$ for some $n$, thus from (*) we get $f(n+1)=-1$ which is impossible. Therefore $0 \in R_{0} \backslash R_{3}=S_{1} \cup S_{2} \cup S_{3}$, thus $k \geqslant 1$.

Next, let us describe the elements $b$ of $R_{0} \backslash R_{3}=S_{1} \cup S_{2} \cup S_{3}$. We claim that each such element satisfies at least one of three conditions (i) $b=0$, (ii) $b=f(0)+1$, and (iii) $b-1 \in S_{1}$. Otherwise $b-1 \in \mathbb{Z}_{\geqslant 0}$, and there exists some $n>0$ such that $f(n)=b-1$; but then $f^{3}(n-1)=f(n)+1=b$, so $b \in R_{3}$.

This yields

$$
3 k=\left|S_{1} \cup S_{2} \cup S_{3}\right| \leqslant 1+1+\left|S_{1}\right|=k+2
$$

or $k \leqslant 1$. Therefore $k=1$, and the inequality above comes to equality. So we have $S_{1}=\{a\}$, $S_{2}=\{f(a)\}$, and $S_{3}=\left\{f^{2}(a)\right\}$ for some $a \in \mathbb{Z}_{\geqslant 0}$, and each one of the three options (i), (ii), and (iii) should be realized exactly once, which means that

$$
\begin{equation*}
\left\{a, f(a), f^{2}(a)\right\}=\{0, a+1, f(0)+1\} \tag{3}
\end{equation*}
$$

III. From (3), we get $a+1 \in\left\{f(a), f^{2}(a)\right\}$ (the case $a+1=a$ is impossible). If $a+1=f^{2}(a)$ then we have $f(a+1)=f^{3}(a)=f(a+1)+1$ which is absurd. Therefore

$$
\begin{equation*}
f(a)=a+1 \tag{4}
\end{equation*}
$$

Next, again from (3) we have $0 \in\left\{a, f^{2}(a)\right\}$. Let us consider these two cases separately. Case 1. Assume that $a=0$, then $f(0)=f(a)=a+1=1$. Also from (3) we get $f(1)=f^{2}(a)=$ $f(0)+1=2$. Now, let us show that $f(n)=n+1$ by induction on $n$; the base cases $n \leqslant 1$ are established. Next, if $n \geqslant 2$ then the induction hypothesis implies

$$
n+1=f(n-1)+1=f^{3}(n-2)=f^{2}(n-1)=f(n),
$$

establishing the step. In this case we have obtained the first of two answers; checking that is satisfies (*) is straightforward.
Case 2. Assume now that $f^{2}(a)=0$; then by (3) we get $a=f(0)+1$. By (4) we get $f(a+1)=$ $f^{2}(a)=0$, then $f(0)=f^{3}(a)=f(a+1)+1=1$, hence $a=f(0)+1=2$ and $f(2)=3$ by (4). To summarize,

$$
f(0)=1, \quad f(2)=3, \quad f(3)=0
$$

Now let us prove by induction on $m$ that (1) holds for all $n=4 k, 4 k+2,4 k+3$ with $k \leqslant m$ and for all $n=4 k+1$ with $k<m$. The base case $m=0$ is established above. For the step, assume that $m \geqslant 1$. From $(*)$ we get $f^{3}(4 m-3)=f(4 m-2)+1=4 m$. Next, by ( 2 ) we have

$$
f(4 m)=f^{4}(4 m-3)=f^{4}(4 m-4)+1=f^{3}(4 m-3)+1=4 m+1
$$

Then by the induction hypothesis together with (*) we successively obtain

$$
\begin{aligned}
& f(4 m-3)=f^{3}(4 m-1)=f(4 m)+1=4 m+2, \\
& f(4 m+2)=f^{3}(4 m-4)=f(4 m-3)+1=4 m+3, \\
& f(4 m+3)=f^{3}(4 m-3)=f(4 m-2)+1=4 m
\end{aligned}
$$

thus finishing the induction step.
Finally, it is straightforward to check that the constructed function works:

$$
\begin{aligned}
f^{3}(4 k) & =4 k+7=f(4 k+1)+1, & & f^{3}(4 k+1)
\end{aligned}=4 k+4=f(4 k+2)+1, ~ 子 r y(4 k+4)+1 .
$$

Solution 2. I. For convenience, let us introduce the function $g(n)=f(n)+1$. Substituting $f(n)$ instead of $n$ into (*) we obtain

$$
\begin{equation*}
f^{4}(n)=f(f(n)+1)+1, \quad \text { or } \quad f^{4}(n)=g^{2}(n) . \tag{5}
\end{equation*}
$$

Applying $f$ to both parts of (*) and using (5) we get

$$
\begin{equation*}
f^{4}(n)+1=f(f(n+1)+1)+1=f^{4}(n+1) \tag{6}
\end{equation*}
$$

Thus, if $g^{2}(0)=f^{4}(0)=c$ then an easy induction on $n$ shows that

$$
\begin{equation*}
g^{2}(n)=f^{4}(n)=n+c, \quad n \in \mathbb{Z}_{\geqslant 0} . \tag{7}
\end{equation*}
$$

This relation implies that both $f$ and $g$ are injective: if, say, $f(m)=f(n)$ then $m+c=$ $f^{4}(m)=f^{4}(n)=n+c$. Next, since $g(n) \geqslant 1$ for every $n$, we have $c=g^{2}(0) \geqslant 1$. Thus from (7) again we obtain $f(n) \neq n$ and $g(n) \neq n$ for all $n \in \mathbb{Z}_{\geqslant 0}$.
II. Next, application of $f$ and $g$ to (7) yields

$$
\begin{equation*}
f(n+c)=f^{5}(n)=f^{4}(f(n))=f(n)+c \quad \text { and } \quad g(n+c)=g^{3}(n)=g(n)+c \tag{8}
\end{equation*}
$$

In particular, this means that if $m \equiv n(\bmod c)$ then $f(m) \equiv f(n)(\bmod c)$. Conversely, if $f(m) \equiv f(n)(\bmod c)$ then we get $m+c=f^{4}(m) \equiv f^{4}(n)=n+c(\bmod c)$. Thus,

$$
\begin{equation*}
m \equiv n \quad(\bmod c) \Longleftrightarrow f(m) \equiv f(n) \quad(\bmod c) \Longleftrightarrow g(m) \equiv g(n) \quad(\bmod c) \tag{9}
\end{equation*}
$$

Now, let us introduce the function $\delta(n)=f(n)-n=g(n)-n-1$. Set

$$
S=\sum_{n=0}^{c-1} \delta(n)
$$

Using (8), we get that for every complete residue system $n_{1}, \ldots, n_{c}$ modulo $c$ we also have

$$
S=\sum_{i=1}^{c} \delta\left(n_{i}\right)
$$

By (9), we get that $\left\{f^{k}(n): n=0, \ldots, c-1\right\}$ and $\left\{g^{k}(n): n=0, \ldots, c-1\right\}$ are complete residue systems modulo $c$ for all $k$. Thus we have

$$
c^{2}=\sum_{n=0}^{c-1}\left(f^{4}(n)-n\right)=\sum_{k=0}^{3} \sum_{n=0}^{c-1}\left(f^{k+1}(n)-f^{k}(n)\right)=\sum_{k=0}^{3} \sum_{n=0}^{c-1} \delta\left(f^{k}(n)\right)=4 S
$$

and similarly

$$
c^{2}=\sum_{n=0}^{c-1}\left(g^{2}(n)-n\right)=\sum_{k=0}^{1} \sum_{n=0}^{c-1}\left(g^{k+1}(n)-g^{k}(n)\right)=\sum_{k=0}^{1} \sum_{n=0}^{c-1}\left(\delta\left(g^{k}(n)\right)+1\right)=2 S+2 c .
$$

Therefore $c^{2}=4 S=2 \cdot 2 S=2\left(c^{2}-2 c\right)$, or $c^{2}=4 c$. Since $c \neq 0$, we get $c=4$. Thus, in view of (8) it is sufficient to determine the values of $f$ on the numbers $0,1,2,3$.
III. Let $d=g(0) \geqslant 1$. Then $g(d)=g^{2}(0)=0+c=4$. Now, if $d \geqslant 4$, then we would have $g(d-4)=g(d)-4=0$ which is impossible. Thus $d \in\{1,2,3\}$. If $d=1$ then we have $f(0)=g(0)-1=0$ which is impossible since $f(n) \neq n$ for all $n$. If $d=3$ then $g(3)=g^{2}(0)=4$ and hence $f(3)=3$ which is also impossible. Thus $g(0)=2$ and hence $g(2)=g^{2}(0)=4$.

Next, if $g(1)=1+4 k$ for some integer $k$, then $5=g^{2}(1)=g(1+4 k)=g(1)+4 k=1+8 k$ which is impossible. Thus, since $\{g(n): n=0,1,2,3\}$ is a complete residue system modulo 4 , we get $g(1)=3+4 k$ and hence $g(3)=g^{2}(1)-4 k=5-4 k$, leading to $k=0$ or $k=1$. So, we obtain iether

$$
f(0)=1, f(1)=2, f(2)=3, f(3)=4, \quad \text { or } \quad f(0)=1, f(1)=6, f(2)=3, f(3)=0,
$$

thus arriving to the two functions listed in the answer.
Finally, one can check that these two function work as in Solution 1. One may simplify the checking by noticing that (8) allows us to reduce it to $n=0,1,2,3$.

A6. Let $m \neq 0$ be an integer. Find all polynomials $P(x)$ with real coefficients such that

$$
\begin{equation*}
\left(x^{3}-m x^{2}+1\right) P(x+1)+\left(x^{3}+m x^{2}+1\right) P(x-1)=2\left(x^{3}-m x+1\right) P(x) \tag{1}
\end{equation*}
$$

for all real numbers $x$.
(Serbia)
Answer. $P(x)=t x$ for any real number $t$.
Solution. Let $P(x)=a_{n} x^{n}+\cdots+a_{0} x^{0}$ with $a_{n} \neq 0$. Comparing the coefficients of $x^{n+1}$ on both sides gives $a_{n}(n-2 m)(n-1)=0$, so $n=1$ or $n=2 m$.

If $n=1$, one easily verifies that $P(x)=x$ is a solution, while $P(x)=1$ is not. Since the given condition is linear in $P$, this means that the linear solutions are precisely $P(x)=t x$ for $t \in \mathbb{R}$.

Now assume that $n=2 m$. The polynomial $x P(x+1)-(x+1) P(x)=(n-1) a_{n} x^{n}+\cdots$ has degree $n$, and therefore it has at least one (possibly complex) root $r$. If $r \notin\{0,-1\}$, define $k=P(r) / r=P(r+1) /(r+1)$. If $r=0$, let $k=P(1)$. If $r=-1$, let $k=-P(-1)$. We now consider the polynomial $S(x)=P(x)-k x$. It also satisfies (1) because $P(x)$ and $k x$ satisfy it. Additionally, it has the useful property that $r$ and $r+1$ are roots.

Let $A(x)=x^{3}-m x^{2}+1$ and $B(x)=x^{3}+m x^{2}+1$. Plugging in $x=s$ into (1) implies that:
If $s-1$ and $s$ are roots of $S$ and $s$ is not a root of $A$, then $s+1$ is a root of $S$.
If $s$ and $s+1$ are roots of $S$ and $s$ is not a root of $B$, then $s-1$ is a root of $S$.
Let $a \geqslant 0$ and $b \geqslant 1$ be such that $r-a, r-a+1, \ldots, r, r+1, \ldots, r+b-1, r+b$ are roots of $S$, while $r-a-1$ and $r+b+1$ are not. The two statements above imply that $r-a$ is a root of $B$ and $r+b$ is a root of $A$.

Since $r-a$ is a root of $B(x)$ and of $A(x+a+b)$, it is also a root of their greatest common divisor $C(x)$ as integer polynomials. If $C(x)$ was a non-trivial divisor of $B(x)$, then $B$ would have a rational root $\alpha$. Since the first and last coefficients of $B$ are $1, \alpha$ can only be 1 or -1 ; but $B(-1)=m>0$ and $B(1)=m+2>0$ since $n=2 m$.

Therefore $B(x)=A(x+a+b)$. Writing $c=a+b \geqslant 1$ we compute

$$
0=A(x+c)-B(x)=(3 c-2 m) x^{2}+c(3 c-2 m) x+c^{2}(c-m)
$$

Then we must have $3 c-2 m=c-m=0$, which gives $m=0$, a contradiction. We conclude that $f(x)=t x$ is the only solution.

Solution 2. Multiplying (1) by $x$, we rewrite it as

$$
x\left(x^{3}-m x^{2}+1\right) P(x+1)+x\left(x^{3}+m x^{2}+1\right) P(x-1)=[(x+1)+(x-1)]\left(x^{3}-m x+1\right) P(x) .
$$

After regrouping, it becomes

$$
\begin{equation*}
\left(x^{3}-m x^{2}+1\right) Q(x)=\left(x^{3}+m x^{2}+1\right) Q(x-1) \tag{2}
\end{equation*}
$$

where $Q(x)=x P(x+1)-(x+1) P(x)$. If $\operatorname{deg} P \geqslant 2$ then $\operatorname{deg} Q=\operatorname{deg} P$, so $Q(x)$ has a finite multiset of complex roots, which we denote $R_{Q}$. Each root is taken with its multiplicity. Then the multiset of complex roots of $Q(x-1)$ is $R_{Q}+1=\left\{z+1: z \in R_{Q}\right\}$.

Let $\left\{x_{1}, x_{2}, x_{3}\right\}$ and $\left\{y_{1}, y_{2}, y_{3}\right\}$ be the multisets of roots of the polynomials $A(x)=x^{3}-m x^{2}+1$ and $B(x)=x^{3}+m x^{2}+1$, respectively. From (2) we get the equality of multisets

$$
\left\{x_{1}, x_{2}, x_{3}\right\} \cup R_{Q}=\left\{y_{1}, y_{2}, y_{3}\right\} \cup\left(R_{Q}+1\right)
$$

For every $r \in R_{Q}$, since $r+1$ is in the set of the right hand side, we must have $r+1 \in R_{Q}$ or $r+1=x_{i}$ for some $i$. Similarly, since $r$ is in the set of the left hand side, either $r-1 \in R_{Q}$ or $r=y_{i}$ for some $i$. This implies that, possibly after relabelling $y_{1}, y_{2}, y_{3}$, all the roots of (2) may be partitioned into three chains of the form $\left\{y_{i}, y_{i}+1, \ldots, y_{i}+k_{i}=x_{i}\right\}$ for $i=1,2,3$ and some integers $k_{1}, k_{2}, k_{3} \geqslant 0$.

Now we analyze the roots of the polynomial $A_{a}(x)=x^{3}+a x^{2}+1$. Using calculus or elementary methods, we find that the local extrema of $A_{a}(x)$ occur at $x=0$ and $x=-2 a / 3$; their values are $A_{a}(0)=1>0$ and $A_{a}(-2 a / 3)=1+4 a^{3} / 27$, which is positive for integers $a \geqslant-1$ and negative for integers $a \leqslant-2$. So when $a \in \mathbb{Z}, A_{a}$ has three real roots if $a \leqslant-2$ and one if $a \geqslant-1$.

Now, since $y_{i}-x_{i} \in \mathbb{Z}$ for $i=1,2,3$, the cubics $A_{m}$ and $A_{-m}$ must have the same number of real roots. The previous analysis then implies that $m=1$ or $m=-1$. Therefore the real root $\alpha$ of $A_{1}(x)=x^{3}+x^{2}+1$ and the real root $\beta$ of $A_{-1}(x)=x^{3}-x^{2}+1$ must differ by an integer. But this is impossible, because $A_{1}\left(-\frac{3}{2}\right)=-\frac{1}{8}$ and $A_{1}(-1)=1$ so $-1.5<\alpha<-1$, while $A_{-1}(-1)=-1$ and $A_{-1}\left(-\frac{1}{2}\right)=\frac{5}{8}$, so $-1<\beta<-0.5$.

It follows that $\operatorname{deg} P \leqslant 1$. Then, as shown in Solution 1, we conclude that the solutions are $P(x)=t x$ for all real numbers $t$.

## Combinatorics

C1. Let $n$ be a positive integer. Find the smallest integer $k$ with the following property: Given any real numbers $a_{1}, \ldots, a_{d}$ such that $a_{1}+a_{2}+\cdots+a_{d}=n$ and $0 \leqslant a_{i} \leqslant 1$ for $i=1,2, \ldots, d$, it is possible to partition these numbers into $k$ groups (some of which may be empty) such that the sum of the numbers in each group is at most 1 .
(Poland)
Answer. $k=2 n-1$.
Solution 1. If $d=2 n-1$ and $a_{1}=\cdots=a_{2 n-1}=n /(2 n-1)$, then each group in such a partition can contain at most one number, since $2 n /(2 n-1)>1$. Therefore $k \geqslant 2 n-1$. It remains to show that a suitable partition into $2 n-1$ groups always exists.

We proceed by induction on $d$. For $d \leqslant 2 n-1$ the result is trivial. If $d \geqslant 2 n$, then since

$$
\left(a_{1}+a_{2}\right)+\ldots+\left(a_{2 n-1}+a_{2 n}\right) \leqslant n
$$

we may find two numbers $a_{i}, a_{i+1}$ such that $a_{i}+a_{i+1} \leqslant 1$. We "merge" these two numbers into one new number $a_{i}+a_{i+1}$. By the induction hypothesis, a suitable partition exists for the $d-1$ numbers $a_{1}, \ldots, a_{i-1}, a_{i}+a_{i+1}, a_{i+2}, \ldots, a_{d}$. This induces a suitable partition for $a_{1}, \ldots, a_{d}$.

Solution 2. We will show that it is even possible to split the sequence $a_{1}, \ldots, a_{d}$ into $2 n-1$ contiguous groups so that the sum of the numbers in each groups does not exceed 1. Consider a segment $S$ of length $n$, and partition it into segments $S_{1}, \ldots, S_{d}$ of lengths $a_{1}, \ldots, a_{d}$, respectively, as shown below. Consider a second partition of $S$ into $n$ equal parts by $n-1$ "empty dots".


Assume that the $n-1$ empty dots are in segments $S_{i_{1}}, \ldots, S_{i_{n-1}}$. (If a dot is on the boundary of two segments, we choose the right segment). These $n-1$ segments are distinct because they have length at most 1. Consider the partition:

$$
\left\{a_{1}, \ldots, a_{i_{1}-1}\right\},\left\{a_{i_{1}}\right\},\left\{a_{i_{1}+1}, \ldots, a_{i_{2}-1}\right\},\left\{a_{i_{2}}\right\}, \ldots\left\{a_{i_{n-1}}\right\},\left\{a_{i_{n-1}+1}, \ldots, a_{d}\right\} .
$$

In the example above, this partition is $\left\{a_{1}, a_{2}\right\},\left\{a_{3}\right\},\left\{a_{4}, a_{5}\right\},\left\{a_{6}\right\}, \varnothing,\left\{a_{7}\right\},\left\{a_{8}, a_{9}, a_{10}\right\}$. We claim that in this partition, the sum of the numbers in this group is at most 1 .

For the sets $\left\{a_{i_{t}}\right\}$ this is obvious since $a_{i_{t}} \leqslant 1$. For the sets $\left\{a_{i_{t}}+1, \ldots, a_{i_{t+1}-1}\right\}$ this follows from the fact that the corresponding segments lie between two neighboring empty dots, or between an endpoint of $S$ and its nearest empty dot. Therefore the sum of their lengths cannot exceed 1.

Solution 3. First put all numbers greater than $\frac{1}{2}$ in their own groups. Then, form the remaining groups as follows: For each group, add new $a_{i}$ s one at a time until their sum exceeds $\frac{1}{2}$. Since the last summand is at most $\frac{1}{2}$, this group has sum at most 1 . Continue this procedure until we have used all the $a_{i}$ s. Notice that the last group may have sum less than $\frac{1}{2}$. If the sum of the numbers in the last two groups is less than or equal to 1 , we merge them into one group. In the end we are left with $m$ groups. If $m=1$ we are done. Otherwise the first $m-2$ have sums greater than $\frac{1}{2}$ and the last two have total sum greater than 1 . Therefore $n>(m-2) / 2+1$ so $m \leqslant 2 n-1$ as desired.

Comment 1. The original proposal asked for the minimal value of $k$ when $n=2$.
Comment 2. More generally, one may ask the same question for real numbers between 0 and 1 whose sum is a real number $r$. In this case the smallest value of $k$ is $k=\lceil 2 r\rceil-1$, as Solution 3 shows.

Solutions 1 and 2 lead to the slightly weaker bound $k \leqslant 2\lceil r\rceil-1$. This is actually the optimal bound for partitions into consecutive groups, which are the ones contemplated in these two solutions. To see this, assume that $r$ is not an integer and let $c=(r+1-\lceil r\rceil) /(1+\lceil r\rceil)$. One easily checks that $0<c<\frac{1}{2}$ and $\lceil r\rceil(2 c)+(\lceil r\rceil-1)(1-c)=r$, so the sequence

$$
2 c, 1-c, 2 c, 1-c, \ldots, 1-c, 2 c
$$

of $2\lceil r\rceil-1$ numbers satisfies the given conditions. For this sequence, the only suitable partition into consecutive groups is the trivial partition, which requires $2\lceil r\rceil-1$ groups.

C2. In the plane, 2013 red points and 2014 blue points are marked so that no three of the marked points are collinear. One needs to draw $k$ lines not passing through the marked points and dividing the plane into several regions. The goal is to do it in such a way that no region contains points of both colors.

Find the minimal value of $k$ such that the goal is attainable for every possible configuration of 4027 points.
(Australia)
Answer. $k=2013$.
Solution 1. Firstly, let us present an example showing that $k \geqslant 2013$. Mark 2013 red and 2013 blue points on some circle alternately, and mark one more blue point somewhere in the plane. The circle is thus split into 4026 arcs, each arc having endpoints of different colors. Thus, if the goal is reached, then each arc should intersect some of the drawn lines. Since any line contains at most two points of the circle, one needs at least 4026/2 $=2013$ lines.

It remains to prove that one can reach the goal using 2013 lines. First of all, let us mention that for every two points $A$ and $B$ having the same color, one can draw two lines separating these points from all other ones. Namely, it suffices to take two lines parallel to $A B$ and lying on different sides of $A B$ sufficiently close to it: the only two points between these lines will be $A$ and $B$.

Now, let $P$ be the convex hull of all marked points. Two cases are possible.
Case 1. Assume that $P$ has a red vertex $A$. Then one may draw a line separating $A$ from all the other points, pair up the other 2012 red points into 1006 pairs, and separate each pair from the other points by two lines. Thus, 2013 lines will be used.
Case 2. Assume now that all the vertices of $P$ are blue. Consider any two consecutive vertices of $P$, say $A$ and $B$. One may separate these two points from the others by a line parallel to $A B$. Then, as in the previous case, one pairs up all the other 2012 blue points into 1006 pairs, and separates each pair from the other points by two lines. Again, 2013 lines will be used.

Comment 1. Instead of considering the convex hull, one may simply take a line containing two marked points $A$ and $B$ such that all the other marked points are on one side of this line. If one of $A$ and $B$ is red, then one may act as in Case 1; otherwise both are blue, and one may act as in Case 2.
Solution 2. Let us present a different proof of the fact that $k=2013$ suffices. In fact, we will prove a more general statement:

If $n$ points in the plane, no three of which are collinear, are colored in red and blue arbitrarily, then it suffices to draw $\lfloor n / 2\rfloor$ lines to reach the goal.

We proceed by induction on $n$. If $n \leqslant 2$ then the statement is obvious. Now assume that $n \geqslant 3$, and consider a line $\ell$ containing two marked points $A$ and $B$ such that all the other marked points are on one side of $\ell$; for instance, any line containing a side of the convex hull works.

Remove for a moment the points $A$ and $B$. By the induction hypothesis, for the remaining configuration it suffices to draw $\lfloor n / 2\rfloor-1$ lines to reach the goal. Now return the points $A$ and $B$ back. Three cases are possible.
Case 1. If $A$ and $B$ have the same color, then one may draw a line parallel to $\ell$ and separating $A$ and $B$ from the other points. Obviously, the obtained configuration of $\lfloor n / 2\rfloor$ lines works.
Case 2. If $A$ and $B$ have different colors, but they are separated by some drawn line, then again the same line parallel to $\ell$ works.

Case 3. Finally, assume that $A$ and $B$ have different colors and lie in one of the regions defined by the drawn lines. By the induction assumption, this region contains no other points of one of the colors - without loss of generality, the only blue point it contains is $A$. Then it suffices to draw a line separating $A$ from all other points.

Thus the step of the induction is proved.

Comment 2. One may ask a more general question, replacing the numbers 2013 and 2014 by any positive integers $m$ and $n$, say with $m \leqslant n$. Denote the answer for this problem by $f(m, n)$.

One may show along the lines of Solution 1 that $m \leqslant f(m, n) \leqslant m+1$; moreover, if $m$ is even then $f(m, n)=m$. On the other hand, for every odd $m$ there exists an $N$ such that $f(m, n)=m$ for all $m \leqslant n \leqslant N$, and $f(m, n)=m+1$ for all $n>N$.

C3. A crazy physicist discovered a new kind of particle which he called an imon, after some of them mysteriously appeared in his lab. Some pairs of imons in the lab can be entangled, and each imon can participate in many entanglement relations. The physicist has found a way to perform the following two kinds of operations with these particles, one operation at a time.
(i) If some imon is entangled with an odd number of other imons in the lab, then the physicist can destroy it.
(ii) At any moment, he may double the whole family of imons in his lab by creating a copy $I^{\prime}$ of each imon $I$. During this procedure, the two copies $I^{\prime}$ and $J^{\prime}$ become entangled if and only if the original imons $I$ and $J$ are entangled, and each copy $I^{\prime}$ becomes entangled with its original imon $I$; no other entanglements occur or disappear at this moment.

Prove that the physicist may apply a sequence of such operations resulting in a family of imons, no two of which are entangled.
(Japan)
Solution 1. Let us consider a graph with the imons as vertices, and two imons being connected if and only if they are entangled. Recall that a proper coloring of a graph $G$ is a coloring of its vertices in several colors so that every two connected vertices have different colors.
Lemma. Assume that a graph $G$ admits a proper coloring in $n$ colors $(n>1)$. Then one may perform a sequence of operations resulting in a graph which admits a proper coloring in $n-1$ colors.
Proof. Let us apply repeatedly operation (i) to any appropriate vertices while it is possible. Since the number of vertices decreases, this process finally results in a graph where all the degrees are even. Surely this graph also admits a proper coloring in $n$ colors $1, \ldots, n$; let us fix this coloring.

Now apply the operation (ii) to this graph. A proper coloring of the resulting graph in $n$ colors still exists: one may preserve the colors of the original vertices and color the vertex $I^{\prime}$ in a color $k+1(\bmod n)$ if the vertex $I$ has color $k$. Then two connected original vertices still have different colors, and so do their two connected copies. On the other hand, the vertices $I$ and $I^{\prime}$ have different colors since $n>1$.

All the degrees of the vertices in the resulting graph are odd, so one may apply operation $(i)$ to delete consecutively all the vertices of color $n$ one by one; no two of them are connected by an edge, so their degrees do not change during the process. Thus, we obtain a graph admitting a proper coloring in $n-1$ colors, as required. The lemma is proved.

Now, assume that a graph $G$ has $n$ vertices; then it admits a proper coloring in $n$ colors. Applying repeatedly the lemma we finally obtain a graph admitting a proper coloring in one color, that is - a graph with no edges, as required.

Solution 2. Again, we will use the graph language.
I. We start with the following observation.

Lemma. Assume that a graph $G$ contains an isolated vertex $A$, and a graph $G^{\circ}$ is obtained from $G$ by deleting this vertex. Then, if one can apply a sequence of operations which makes a graph with no edges from $G^{\circ}$, then such a sequence also exists for $G$.
Proof. Consider any operation applicable to $G^{\circ}$ resulting in a graph $G_{1}^{\circ}$; then there exists a sequence of operations applicable to $G$ and resulting in a graph $G_{1}$ differing from $G_{1}^{\circ}$ by an addition of an isolated vertex $A$. Indeed, if this operation is of type $(i)$, then one may simply repeat it in $G$.

Otherwise, the operation is of type (ii), and one may apply it to $G$ and then delete the vertex $A^{\prime}$ (it will have degree 1).

Thus one may change the process for $G^{\circ}$ into a corresponding process for $G$ step by step.
In view of this lemma, if at some moment a graph contains some isolated vertex, then we may simply delete it; let us call this operation (iii).
II. Let $V=\left\{A_{1}^{0}, \ldots, A_{n}^{0}\right\}$ be the vertices of the initial graph. Let us describe which graphs can appear during our operations. Assume that operation (ii) was applied $m$ times. If these were the only operations applied, then the resulting graph $G_{n}^{m}$ has the set of vertices which can be enumerated as

$$
V_{n}^{m}=\left\{A_{i}^{j}: 1 \leqslant i \leqslant n, 0 \leqslant j \leqslant 2^{m}-1\right\},
$$

where $A_{i}^{0}$ is the common "ancestor" of all the vertices $A_{i}^{j}$, and the binary expansion of $j$ (adjoined with some zeroes at the left to have $m$ digits) "keeps the history" of this vertex: the $d$ th digit from the right is 0 if at the $d$ th doubling the ancestor of $A_{i}^{j}$ was in the original part, and this digit is 1 if it was in the copy.

Next, the two vertices $A_{i}^{j}$ and $A_{k}^{\ell}$ in $G_{n}^{m}$ are connected with an edge exactly if either (1) $j=\ell$ and there was an edge between $A_{i}^{0}$ and $A_{k}^{0}$ (so these vertices appeared at the same application of operation (ii)); or (2) $i=k$ and the binary expansions of $j$ and $\ell$ differ in exactly one digit (so their ancestors became connected as a copy and the original vertex at some application of (ii)).

Now, if some operations $(i)$ were applied during the process, then simply some vertices in $G_{n}^{m}$ disappeared. So, in any case the resulting graph is some induced subgraph of $G_{n}^{m}$.
III. Finally, we will show that from each (not necessarily induced) subgraph of $G_{n}^{m}$ one can obtain a graph with no vertices by applying operations $(i)$, (ii) and (iii). We proceed by induction on $n$; the base case $n=0$ is trivial.

For the induction step, let us show how to apply several operations so as to obtain a graph containing no vertices of the form $A_{n}^{j}$ for $j \in \mathbb{Z}$. We will do this in three steps.
Step 1. We apply repeatedly operation $(i)$ to any appropriate vertices while it is possible. In the resulting graph, all vertices have even degrees.
Step 2. Apply operation (ii) obtaining a subgraph of $G_{n}^{m+1}$ with all degrees being odd. In this graph, we delete one by one all the vertices $A_{n}^{j}$ where the sum of the binary digits of $j$ is even; it is possible since there are no edges between such vertices, so all their degrees remain odd. After that, we delete all isolated vertices.
Step 3. Finally, consider any remaining vertex $A_{n}^{j}$ (then the sum of digits of $j$ is odd). If its degree is odd, then we simply delete it. Otherwise, since $A_{n}^{j}$ is not isolated, we consider any vertex adjacent to it. It has the form $A_{k}^{j}$ for some $k<n$ (otherwise it would have the form $A_{n}^{\ell}$, where $\ell$ has an even digit sum; but any such vertex has already been deleted at Step 2). No neighbor of $A_{k}^{j}$ was deleted at Steps 2 and 3, so it has an odd degree. Then we successively delete $A_{k}^{j}$ and $A_{n}^{j}$.

Notice that this deletion does not affect the applicability of this step to other vertices, since no two vertices $A_{i}^{j}$ and $A_{k}^{\ell}$ for different $j, \ell$ with odd digit sum are connected with an edge. Thus we will delete all the remaining vertices of the form $A_{n}^{j}$, obtaining a subgraph of $G_{n-1}^{m+1}$. The application of the induction hypothesis finishes the proof.

Comment. In fact, the graph $G_{n}^{m}$ is a Cartesian product of $G$ and the graph of an $m$-dimensional hypercube.

C4. Let $n$ be a positive integer, and let $A$ be a subset of $\{1, \ldots, n\}$. An $A$-partition of $n$ into $k$ parts is a representation of $n$ as a sum $n=a_{1}+\cdots+a_{k}$, where the parts $a_{1}, \ldots, a_{k}$ belong to $A$ and are not necessarily distinct. The number of different parts in such a partition is the number of (distinct) elements in the set $\left\{a_{1}, a_{2}, \ldots, a_{k}\right\}$.

We say that an $A$-partition of $n$ into $k$ parts is optimal if there is no $A$-partition of $n$ into $r$ parts with $r<k$. Prove that any optimal $A$-partition of $n$ contains at most $\sqrt[3]{6 n}$ different parts.
(Germany)
Solution 1. If there are no $A$-partitions of $n$, the result is vacuously true. Otherwise, let $k_{\text {min }}$ be the minimum number of parts in an $A$-partition of $n$, and let $n=a_{1}+\cdots+a_{k_{\min }}$ be an optimal partition. Denote by $s$ the number of different parts in this partition, so we can write $S=\left\{a_{1}, \ldots, a_{k_{\min }}\right\}=\left\{b_{1}, \ldots, b_{s}\right\}$ for some pairwise different numbers $b_{1}<\cdots<b_{s}$ in $A$.

If $s>\sqrt[3]{6 n}$, we will prove that there exist subsets $X$ and $Y$ of $S$ such that $|X|<|Y|$ and $\sum_{x \in X} x=\sum_{y \in Y} y$. Then, deleting the elements of $Y$ from our partition and adding the elements of $X$ to it, we obtain an $A$-partition of $n$ into less than $k_{\text {min }}$ parts, which is the desired contradiction.

For each positive integer $k \leqslant s$, we consider the $k$-element subset

$$
S_{1,0}^{k}:=\left\{b_{1}, \ldots, b_{k}\right\}
$$

as well as the following $k$-element subsets $S_{i, j}^{k}$ of $S$ :

$$
S_{i, j}^{k}:=\left\{b_{1}, \ldots, b_{k-i}, b_{k-i+j+1}, b_{s-i+2}, \ldots, b_{s}\right\}, \quad i=1, \ldots, k, \quad j=1, \ldots, s-k .
$$

Pictorially, if we represent the elements of $S$ by a sequence of dots in increasing order, and represent a subset of $S$ by shading in the appropriate dots, we have:


Denote by $\Sigma_{i, j}^{k}$ the sum of elements in $S_{i, j}^{k}$. Clearly, $\Sigma_{1,0}^{k}$ is the minimum sum of a $k$-element subset of $S$. Next, for all appropriate indices $i$ and $j$ we have

$$
\sum_{i, j}^{k}=\sum_{i, j+1}^{k}+b_{k-i+j+1}-b_{k-i+j+2}<\sum_{i, j+1}^{k} \quad \text { and } \quad \sum_{i, s-k}^{k}=\sum_{i+1,1}^{k}+b_{k-i}-b_{k-i+1}<\sum_{i+1,1}^{k} .
$$

Therefore

$$
1 \leqslant \Sigma_{1,0}^{k}<\Sigma_{1,1}^{k}<\sum_{1,2}^{k}<\cdots<\Sigma_{1, s-k}^{k}<\sum_{2,1}^{k}<\cdots<\sum_{2, s-k}^{k}<\Sigma_{3,1}^{k}<\cdots<\Sigma_{k, s-k}^{k} \leqslant n
$$

To see this in the picture, we start with the $k$ leftmost points marked. At each step, we look for the rightmost point which can move to the right, and move it one unit to the right. We continue until the $k$ rightmost points are marked. As we do this, the corresponding sums clearly increase.

For each $k$ we have found $k(s-k)+1$ different integers of the form $\Sigma_{i, j}^{k}$ between 1 and $n$. As we vary $k$, the total number of integers we are considering is

$$
\sum_{k=1}^{s}(k(s-k)+1)=s \cdot \frac{s(s+1)}{2}-\frac{s(s+1)(2 s+1)}{6}+s=\frac{s\left(s^{2}+5\right)}{6}>\frac{s^{3}}{6}>n .
$$

Since they are between 1 and $n$, at least two of these integers are equal. Consequently, there exist $1 \leqslant k<k^{\prime} \leqslant s$ and $X=S_{i, j}^{k}$ as well as $Y=S_{i^{\prime}, j^{\prime}}^{k^{\prime}}$ such that

$$
\sum_{x \in X} x=\sum_{y \in Y} y, \quad \text { but } \quad|X|=k<k^{\prime}=|Y|,
$$

as required. The result follows.

Solution 2. Assume, to the contrary, that the statement is false, and choose the minimum number $n$ for which it fails. So there exists a set $A \subseteq\{1, \ldots, n\}$ together with an optimal $A$ partition $n=a_{1}+\cdots+a_{k_{\min }}$ of $n$ refuting our statement, where, of course, $k_{\min }$ is the minimum number of parts in an $A$-partition of $n$. Again, we define $S=\left\{a_{1}, \ldots, a_{k_{\min }}\right\}=\left\{b_{1}, \ldots, b_{s}\right\}$ with $b_{1}<\cdots<b_{s}$; by our assumption we have $s>\sqrt[3]{6 n}>1$. Without loss of generality we assume that $a_{k_{\text {min }}}=b_{s}$. Let us distinguish two cases.
Case 1. $b_{s} \geqslant \frac{s(s-1)}{2}+1$.
Consider the partition $n-b_{s}=a_{1}+\cdots+a_{k_{\min }-1}$, which is clearly a minimum $A$-partition of $n-b_{s}$ with at least $s-1 \geqslant 1$ different parts. Now, from $n<\frac{s^{3}}{6}$ we obtain

$$
n-b_{s} \leqslant n-\frac{s(s-1)}{2}-1<\frac{s^{3}}{6}-\frac{s(s-1)}{2}-1<\frac{(s-1)^{3}}{6}
$$

so $s-1>\sqrt[3]{6\left(n-b_{s}\right)}$, which contradicts the choice of $n$. Case 2. $b_{s} \leqslant \frac{s(s-1)}{2}$.

Set $b_{0}=0, \Sigma_{0,0}=0$, and $\Sigma_{i, j}=b_{1}+\cdots+b_{i-1}+b_{j}$ for $1 \leqslant i \leqslant j<s$. There are $\frac{s(s-1)}{2}+1>b_{s}$ such sums; so at least two of them, say $\Sigma_{i, j}$ and $\Sigma_{i^{\prime}, j^{\prime}}$, are congruent modulo $b_{s}$ (where $(i, j) \neq\left(i^{\prime}, j^{\prime}\right)$ ). This means that $\Sigma_{i, j}-\Sigma_{i^{\prime}, j^{\prime}}=r b_{s}$ for some integer $r$. Notice that for $i \leqslant j<k<s$ we have

$$
0<\Sigma_{i, k}-\Sigma_{i, j}=b_{k}-b_{j}<b_{s}
$$

so the indices $i$ and $i^{\prime}$ are distinct, and we may assume that $i>i^{\prime}$. Next, we observe that $\Sigma_{i, j}-\Sigma_{i^{\prime}, j^{\prime}}=\left(b_{i^{\prime}}-b_{j^{\prime}}\right)+b_{j}+b_{i^{\prime}+1}+\cdots+b_{i-1}$ and $b_{i^{\prime}} \leqslant b_{j^{\prime}}$ imply

$$
-b_{s}<-b_{j^{\prime}}<\Sigma_{i, j}-\sum_{i^{\prime}, j^{\prime}}<\left(i-i^{\prime}\right) b_{s}
$$

so $0 \leqslant r \leqslant i-i^{\prime}-1$.
Thus, we may remove the $i$ terms of $\Sigma_{i, j}$ in our $A$-partition, and replace them by the $i^{\prime}$ terms of $\Sigma_{i^{\prime}, j^{\prime}}$ and $r$ terms equal to $b_{s}$, for a total of $r+i^{\prime}<i$ terms. The result is an $A$-partition of $n$ into a smaller number of parts, a contradiction.

Comment. The original proposal also contained a second part, showing that the estimate appearing in the problem has the correct order of magnitude:
For every positive integer $n$, there exist a set $A$ and an optimal $A$-partition of $n$ that contains $\lfloor\sqrt[3]{2 n}\rfloor$ different parts.

The Problem Selection Committee removed this statement from the problem, since it seems to be less suitable for the competiton; but for completeness we provide an outline of its proof here.

Let $k=\lfloor\sqrt[3]{2 n}\rfloor-1$. The statement is trivial for $n<4$, so we assume $n \geqslant 4$ and hence $k \geqslant 1$. Let $h=\left\lfloor\frac{n-1}{k}\right\rfloor$. Notice that $h \geqslant \frac{n}{k}-1$.

Now let $A=\{1, \ldots, h\}$, and set $a_{1}=h, a_{2}=h-1, \ldots, a_{k}=h-k+1$, and $a_{k+1}=n-\left(a_{1}+\cdots+a_{k}\right)$. It is not difficult to prove that $a_{k}>a_{k+1} \geqslant 1$, which shows that

$$
n=a_{1}+\ldots+a_{k+1}
$$

is an $A$-partition of $n$ into $k+1$ different parts. Since $k h<n$, any $A$-partition of $n$ has at least $k+1$ parts. Therefore our $A$-partition is optimal, and it has $\lfloor\sqrt[3]{2 n}\rfloor$ distinct parts, as desired.

C5. Let $r$ be a positive integer, and let $a_{0}, a_{1}, \ldots$ be an infinite sequence of real numbers. Assume that for all nonnegative integers $m$ and $s$ there exists a positive integer $n \in[m+1, m+r]$ such that

$$
a_{m}+a_{m+1}+\cdots+a_{m+s}=a_{n}+a_{n+1}+\cdots+a_{n+s} .
$$

Prove that the sequence is periodic, i. e. there exists some $p \geqslant 1$ such that $a_{n+p}=a_{n}$ for all $n \geqslant 0$.
(India)
Solution. For every indices $m \leqslant n$ we will denote $S(m, n)=a_{m}+a_{m+1}+\cdots+a_{n-1}$; thus $S(n, n)=0$. Let us start with the following lemma.
Lemma. Let $b_{0}, b_{1}, \ldots$ be an infinite sequence. Assume that for every nonnegative integer $m$ there exists a nonnegative integer $n \in[m+1, m+r]$ such that $b_{m}=b_{n}$. Then for every indices $k \leqslant \ell$ there exists an index $t \in[\ell, \ell+r-1]$ such that $b_{t}=b_{k}$. Moreover, there are at most $r$ distinct numbers among the terms of $\left(b_{i}\right)$.
Proof. To prove the first claim, let us notice that there exists an infinite sequence of indices $k_{1}=k, k_{2}, k_{3}, \ldots$ such that $b_{k_{1}}=b_{k_{2}}=\cdots=b_{k}$ and $k_{i}<k_{i+1} \leqslant k_{i}+r$ for all $i \geqslant 1$. This sequence is unbounded from above, thus it hits each segment of the form $[\ell, \ell+r-1]$ with $\ell \geqslant k$, as required.

To prove the second claim, assume, to the contrary, that there exist $r+1$ distinct numbers $b_{i_{1}}, \ldots, b_{i_{r+1}}$. Let us apply the first claim to $k=i_{1}, \ldots, i_{r+1}$ and $\ell=\max \left\{i_{1}, \ldots, i_{r+1}\right\}$; we obtain that for every $j \in\{1, \ldots, r+1\}$ there exists $t_{j} \in[s, s+r-1]$ such that $b_{t_{j}}=b_{i_{j}}$. Thus the segment [ $s, s+r-1$ ] should contain $r+1$ distinct integers, which is absurd.

Setting $s=0$ in the problem condition, we see that the sequence $\left(a_{i}\right)$ satisfies the condition of the lemma, thus it attains at most $r$ distinct values. Denote by $A_{i}$ the ordered $r$-tuple $\left(a_{i}, \ldots, a_{i+r-1}\right)$; then among $A_{i}$ 's there are at most $r^{r}$ distinct tuples, so for every $k \geqslant 0$ two of the tuples $A_{k}, A_{k+1}, \ldots, A_{k+r^{r}}$ are identical. This means that there exists a positive integer $p \leqslant r^{r}$ such that the equality $A_{d}=A_{d+p}$ holds infinitely many times. Let $D$ be the set of indices $d$ satisfying this relation.

Now we claim that $D$ coincides with the set of all nonnegative integers. Since $D$ is unbounded, it suffices to show that $d \in D$ whenever $d+1 \in D$. For that, denote $b_{k}=S(k, p+k)$. The sequence $b_{0}, b_{1}, \ldots$ satisfies the lemma conditions, so there exists an index $t \in[d+1, d+r]$ such that $S(t, t+p)=S(d, d+p)$. This last relation rewrites as $S(d, t)=S(d+p, t+p)$. Since $A_{d+1}=A_{d+p+1}$, we have $S(d+1, t)=S(d+p+1, t+p)$, therefore we obtain

$$
a_{d}=S(d, t)-S(d+1, t)=S(d+p, t+p)-S(d+p+1, t+p)=a_{d+p}
$$

and thus $A_{d}=A_{d+p}$, as required.
Finally, we get $A_{d}=A_{d+p}$ for all $d$, so in particular $a_{d}=a_{d+p}$ for all $d$, QED.
Comment 1. In the present proof, the upper bound for the minimal period length is $r^{r}$. This bound is not sharp; for instance, one may improve it to $(r-1)^{r}$ for $r \geqslant 3$..

On the other hand, this minimal length may happen to be greater than $r$. For instance, it is easy to check that the sequence with period $(3,-3,3,-3,3,-1,-1,-1)$ satisfies the problem condition for $r=7$.

Comment 2. The conclusion remains true even if the problem condition only holds for every $s \geqslant N$ for some positive integer $N$. To show that, one can act as follows. Firstly, the sums of the form $S(i, i+N)$ attain at most $r$ values, as well as the sums of the form $S(i, i+N+1)$. Thus the terms $a_{i}=S(i, i+N+1)-$ $S(i+1, i+N+1)$ attain at most $r^{2}$ distinct values. Then, among the tuples $A_{k}, A_{k+N}, \ldots, A_{k+r^{2 r} N}$ two
are identical, so for some $p \leqslant r^{2 r}$ the set $D=\left\{d: A_{d}=A_{d+N p}\right\}$ is infinite. The further arguments apply almost literally, with $p$ being replaced by $N p$.

After having proved that such a sequence is also necessarily periodic, one may reduce the bound for the minimal period length to $r^{r}$ - essentially by verifying that the sequence satisfies the original version of the condition.

C6. In some country several pairs of cities are connected by direct two-way flights. It is possible to go from any city to any other by a sequence of flights. The distance between two cities is defined to be the least possible number of flights required to go from one of them to the other. It is known that for any city there are at most 100 cities at distance exactly three from it. Prove that there is no city such that more than 2550 other cities have distance exactly four from it.
(Russia)
Solution. Let us denote by $d(a, b)$ the distance between the cities $a$ and $b$, and by

$$
S_{i}(a)=\{c: d(a, c)=i\}
$$

the set of cities at distance exactly $i$ from city $a$.
Assume that for some city $x$ the set $D=S_{4}(x)$ has size at least 2551. Let $A=S_{1}(x)$. A subset $A^{\prime}$ of $A$ is said to be substantial, if every city in $D$ can be reached from $x$ with four flights while passing through some member of $A^{\prime}$; in other terms, every city in $D$ has distance 3 from some member of $A^{\prime}$, or $D \subseteq \bigcup_{a \in A^{\prime}} S_{3}(a)$. For instance, $A$ itself is substantial. Now let us fix some substantial subset $A^{*}$ of $A$ having the minimal cardinality $m=\left|A^{*}\right|$.

Since

$$
m(101-m) \leqslant 50 \cdot 51=2550
$$

there has to be a city $a \in A^{*}$ such that $\left|S_{3}(a) \cap D\right| \geqslant 102-m$. As $\left|S_{3}(a)\right| \leqslant 100$, we obtain that $S_{3}(a)$ may contain at most $100-(102-m)=m-2$ cities $c$ with $d(c, x) \leqslant 3$. Let us denote by $T=\left\{c \in S_{3}(a): d(x, c) \leqslant 3\right\}$ the set of all such cities, so $|T| \leqslant m-2$. Now, to get a contradiction, we will construct $m-1$ distinct elements in $T$, corresponding to $m-1$ elements of the set $A_{a}=A^{*} \backslash\{a\}$.

Firstly, due to the minimality of $A^{*}$, for each $y \in A_{a}$ there exists some city $d_{y} \in D$ which can only be reached with four flights from $x$ by passing through $y$. So, there is a way to get from $x$ to $d_{y}$ along $x-y-b_{y}-c_{y}-d_{y}$ for some cities $b_{y}$ and $c_{y}$; notice that $d\left(x, b_{y}\right)=2$ and $d\left(x, c_{y}\right)=3$ since this path has the minimal possible length.

Now we claim that all $2(m-1)$ cities of the form $b_{y}, c_{y}$ with $y \in A_{a}$ are distinct. Indeed, no $b_{y}$ may coincide with any $c_{z}$ since their distances from $x$ are different. On the other hand, if one had $b_{y}=b_{z}$ for $y \neq z$, then there would exist a path of length 4 from $x$ to $d_{z}$ via $y$, namely $x-y-b_{z}-c_{z}-d_{z}$; this is impossible by the choice of $d_{z}$. Similarly, $c_{y} \neq c_{z}$ for $y \neq z$.

So, it suffices to prove that for every $y \in A_{a}$, one of the cities $b_{y}$ and $c_{y}$ has distance 3 from $a$ (and thus belongs to $T$ ). For that, notice that $d(a, y) \leqslant 2$ due to the path $a-x-y$, while $d\left(a, d_{y}\right) \geqslant d\left(x, d_{y}\right)-d(x, a)=3$. Moreover, $d\left(a, d_{y}\right) \neq 3$ by the choice of $d_{y}$; thus $d\left(a, d_{y}\right)>3$. Finally, in the sequence $d(a, y), d\left(a, b_{y}\right), d\left(a, c_{y}\right), d\left(a, d_{y}\right)$ the neighboring terms differ by at most 1 , the first term is less than 3 , and the last one is greater than 3 ; thus there exists one which is equal to 3 , as required.

Comment 1. The upper bound 2550 is sharp. This can be seen by means of various examples; one of them is the "Roman Empire": it has one capital, called "Rome", that is connected to 51 semicapitals by internally disjoint paths of length 3. Moreover, each of these semicapitals is connected to 50 rural cities by direct flights.

Comment 2. Observe that, under the conditions of the problem, there exists no bound for the size of $S_{1}(x)$ or $S_{2}(x)$.

Comment 3. The numbers 100 and 2550 appearing in the statement of the problem may be replaced by $n$ and $\left\lfloor\frac{(n+1)^{2}}{4}\right\rfloor$ for any positive integer $n$. Still more generally, one can also replace the pair $(3,4)$ of distances under consideration by any pair $(r, s)$ of positive integers satisfying $r<s \leqslant \frac{3}{2} r$.

To adapt the above proof to this situation, one takes $A=S_{s-r}(x)$ and defines the concept of substantiality as before. Then one takes $A^{*}$ to be a minimal substantial subset of $A$, and for each $y \in A^{*}$ one fixes an element $d_{y} \in S_{s}(x)$ which is only reachable from $x$ by a path of length $s$ by passing through $y$. As before, it suffices to show that for distinct $a, y \in A^{*}$ and a path $y=y_{0}-y_{1}-\ldots-y_{r}=d_{y}$, at least one of the cities $y_{0}, \ldots, y_{r-1}$ has distance $r$ from $a$. This can be done as above; the relation $s \leqslant \frac{3}{2} r$ is used here to show that $d\left(a, y_{0}\right) \leqslant r$.

Moreover, the estimate $\left\lfloor\frac{(n+1)^{2}}{4}\right\rfloor$ is also sharp for every positive integer $n$ and every positive integers $r, s$ with $r<s \leqslant \frac{3}{2} r$. This may be shown by an example similar to that in the previous comment.

C7. Let $n \geqslant 2$ be an integer. Consider all circular arrangements of the numbers $0,1, \ldots, n$; the $n+1$ rotations of an arrangement are considered to be equal. A circular arrangement is called beautiful if, for any four distinct numbers $0 \leqslant a, b, c, d \leqslant n$ with $a+c=b+d$, the chord joining numbers $a$ and $c$ does not intersect the chord joining numbers $b$ and $d$.

Let $M$ be the number of beautiful arrangements of $0,1, \ldots, n$. Let $N$ be the number of pairs $(x, y)$ of positive integers such that $x+y \leqslant n$ and $\operatorname{gcd}(x, y)=1$. Prove that

$$
M=N+1
$$

(Russia)
Solution 1. Given a circular arrangement of $[0, n]=\{0,1, \ldots, n\}$, we define a $k$-chord to be a (possibly degenerate) chord whose (possibly equal) endpoints add up to $k$. We say that three chords of a circle are aligned if one of them separates the other two. Say that $m \geqslant 3$ chords are aligned if any three of them are aligned. For instance, in Figure 1, $A, B$, and $C$ are aligned, while $B, C$, and $D$ are not.


Figure 1


Figure 2

Claim. In a beautiful arrangement, the $k$-chords are aligned for any integer $k$.
Proof. We proceed by induction. For $n \leqslant 3$ the statement is trivial. Now let $n \geqslant 4$, and proceed by contradiction. Consider a beautiful arrangement $S$ where the three $k$-chords $A, B, C$ are not aligned. If $n$ is not among the endpoints of $A, B$, and $C$, then by deleting $n$ from $S$ we obtain a beautiful arrangement $S \backslash\{n\}$ of $[0, n-1]$, where $A, B$, and $C$ are aligned by the induction hypothesis. Similarly, if 0 is not among these endpoints, then deleting 0 and decreasing all the numbers by 1 gives a beautiful arrangement $S \backslash\{0\}$ where $A, B$, and $C$ are aligned. Therefore both 0 and $n$ are among the endpoints of these segments. If $x$ and $y$ are their respective partners, we have $n \geqslant 0+x=k=n+y \geqslant n$. Thus 0 and $n$ are the endpoints of one of the chords; say it is $C$.

Let $D$ be the chord formed by the numbers $u$ and $v$ which are adjacent to 0 and $n$ and on the same side of $C$ as $A$ and $B$, as shown in Figure 2. Set $t=u+v$. If we had $t=n$, the $n$-chords $A$, $B$, and $D$ would not be aligned in the beautiful arrangement $S \backslash\{0, n\}$, contradicting the induction hypothesis. If $t<n$, then the $t$-chord from 0 to $t$ cannot intersect $D$, so the chord $C$ separates $t$ and $D$. The chord $E$ from $t$ to $n-t$ does not intersect $C$, so $t$ and $n-t$ are on the same side of $C$. But then the chords $A, B$, and $E$ are not aligned in $S \backslash\{0, n\}$, a contradiction. Finally, the case $t>n$ is equivalent to the case $t<n$ via the beauty-preserving relabelling $x \mapsto n-x$ for $0 \leqslant x \leqslant n$, which sends $t$-chords to $(2 n-t)$-chords. This proves the Claim.

Having established the Claim, we prove the desired result by induction. The case $n=2$ is trivial. Now assume that $n \geqslant 3$. Let $S$ be a beautiful arrangement of $[0, n]$ and delete $n$ to obtain
the beautiful arrangement $T$ of $[0, n-1]$. The $n$-chords of $T$ are aligned, and they contain every point except 0 . Say $T$ is of Type 1 if 0 lies between two of these $n$-chords, and it is of Type 2 otherwise; i.e., if 0 is aligned with these $n$-chords. We will show that each Type 1 arrangement of $[0, n-1]$ arises from a unique arrangement of $[0, n]$, and each Type 2 arrangement of $[0, n-1]$ arises from exactly two beautiful arrangements of $[0, n]$.

If $T$ is of Type 1 , let 0 lie between chords $A$ and $B$. Since the chord from 0 to $n$ must be aligned with $A$ and $B$ in $S, n$ must be on the other arc between $A$ and $B$. Therefore $S$ can be recovered uniquely from $T$. In the other direction, if $T$ is of Type 1 and we insert $n$ as above, then we claim the resulting arrangement $S$ is beautiful. For $0<k<n$, the $k$-chords of $S$ are also $k$-chords of $T$, so they are aligned. Finally, for $n<k<2 n$, notice that the $n$-chords of $S$ are parallel by construction, so there is an antisymmetry axis $\ell$ such that $x$ is symmetric to $n-x$ with respect to $\ell$ for all $x$. If we had two $k$-chords which intersect, then their reflections across $\ell$ would be two ( $2 n-k$ )-chords which intersect, where $0<2 n-k<n$, a contradiction.

If $T$ is of Type 2, there are two possible positions for $n$ in $S$, on either side of 0 . As above, we check that both positions lead to beautiful arrangements of $[0, n]$.

Hence if we let $M_{n}$ be the number of beautiful arrangements of $[0, n]$, and let $L_{n}$ be the number of beautiful arrangements of $[0, n-1]$ of Type 2, we have

$$
M_{n}=\left(M_{n-1}-L_{n-1}\right)+2 L_{n-1}=M_{n-1}+L_{n-1} .
$$

It then remains to show that $L_{n-1}$ is the number of pairs $(x, y)$ of positive integers with $x+y=n$ and $\operatorname{gcd}(x, y)=1$. Since $n \geqslant 3$, this number equals $\varphi(n)=\#\{x: 1 \leqslant x \leqslant n, \operatorname{gcd}(x, n)=1\}$.

To prove this, consider a Type 2 beautiful arrangement of $[0, n-1]$. Label the positions $0, \ldots, n-1(\bmod n)$ clockwise around the circle, so that number 0 is in position 0 . Let $f(i)$ be the number in position $i$; note that $f$ is a permutation of $[0, n-1]$. Let $a$ be the position such that $f(a)=n-1$.

Since the $n$-chords are aligned with 0 , and every point is in an $n$-chord, these chords are all parallel and

$$
f(i)+f(-i)=n \quad \text { for all } i
$$

Similarly, since the $(n-1)$-chords are aligned and every point is in an $(n-1)$-chord, these chords are also parallel and

$$
f(i)+f(a-i)=n-1 \quad \text { for all } i .
$$

Therefore $f(a-i)=f(-i)-1$ for all $i$; and since $f(0)=0$, we get

$$
\begin{equation*}
f(-a k)=k \quad \text { for all } k \tag{1}
\end{equation*}
$$

Recall that this is an equality modulo $n$. Since $f$ is a permutation, we must have $(a, n)=1$. Hence $L_{n-1} \leqslant \varphi(n)$.

To prove equality, it remains to observe that the labeling (1) is beautiful. To see this, consider four numbers $w, x, y, z$ on the circle with $w+y=x+z$. Their positions around the circle satisfy $(-a w)+(-a y)=(-a x)+(-a z)$, which means that the chord from $w$ to $y$ and the chord from $x$ to $z$ are parallel. Thus (1) is beautiful, and by construction it has Type 2. The desired result follows.

Solution 2. Notice that there are exactly $N$ irreducible fractions $f_{1}<\cdots<f_{N}$ in $(0,1)$ whose denominator is at most $n$, since the pair $(x, y)$ with $x+y \leqslant n$ and $(x, y)=1$ corresponds to the fraction $x /(x+y)$. Write $f_{i}=\frac{a_{i}}{b_{i}}$ for $1 \leqslant i \leqslant N$.

We begin by constructing $N+1$ beautiful arrangements. Take any $\alpha \in(0,1)$ which is not one of the above $N$ fractions. Consider a circle of perimeter 1 . Successively mark points $0,1,2, \ldots, n$ where 0 is arbitrary, and the clockwise distance from $i$ to $i+1$ is $\alpha$. The point $k$ will be at clockwise distance $\{k \alpha\}$ from 0 , where $\{r\}$ denotes the fractional part of $r$. Call such a circular arrangement cyclic and denote it by $A(\alpha)$. If the clockwise order of the points is the same in $A\left(\alpha_{1}\right)$ and $A\left(\alpha_{2}\right)$, we regard them as the same circular arrangement. Figure 3 shows the cyclic arrangement $A(3 / 5+\epsilon)$ of $[0,13]$ where $\epsilon>0$ is very small.


Figure 3
If $0 \leqslant a, b, c, d \leqslant n$ satisfy $a+c=b+d$, then $a \alpha+c \alpha=b \alpha+d \alpha$, so the chord from $a$ to $c$ is parallel to the chord from $b$ to $d$ in $A(\alpha)$. Hence in a cyclic arrangement all $k$-chords are parallel. In particular every cyclic arrangement is beautiful.

Next we show that there are exactly $N+1$ distinct cyclic arrangements. To see this, let us see how $A(\alpha)$ changes as we increase $\alpha$ from 0 to 1 . The order of points $p$ and $q$ changes precisely when we cross a value $\alpha=f$ such that $\{p f\}=\{q f\}$; this can only happen if $f$ is one of the $N$ fractions $f_{1}, \ldots, f_{N}$. Therefore there are at most $N+1$ different cyclic arrangements.

To show they are all distinct, recall that $f_{i}=a_{i} / b_{i}$ and let $\epsilon>0$ be a very small number. In the arrangement $A\left(f_{i}+\epsilon\right)$, point $k$ lands at $\frac{k a_{i}\left(\bmod b_{i}\right)}{b_{i}}+k \epsilon$. Therefore the points are grouped into $b_{i}$ clusters next to the points $0, \frac{1}{b_{i}}, \ldots, \frac{b_{i}-1}{b_{i}}$ of the circle. The cluster following $\frac{k}{b_{i}}$ contains the numbers congruent to $k a_{i}^{-1}$ modulo $b_{i}$, listed clockwise in increasing order. It follows that the first number after 0 in $A\left(f_{i}+\epsilon\right)$ is $b_{i}$, and the first number after 0 which is less than $b_{i}$ is $a_{i}^{-1}\left(\bmod b_{i}\right)$, which uniquely determines $a_{i}$. In this way we can recover $f_{i}$ from the cyclic arrangement. Note also that $A\left(f_{i}+\epsilon\right)$ is not the trivial arrangement where we list $0,1, \ldots, n$ in order clockwise. It follows that the $N+1$ cyclic arrangements $A(\epsilon), A\left(f_{1}+\epsilon\right), \ldots, A\left(f_{N}+\epsilon\right)$ are distinct.

Let us record an observation which will be useful later:

$$
\begin{equation*}
\text { if } f_{i}<\alpha<f_{i+1} \text { then } 0 \text { is immediately after } b_{i+1} \text { and before } b_{i} \text { in } A(\alpha) \text {. } \tag{2}
\end{equation*}
$$

Indeed, we already observed that $b_{i}$ is the first number after 0 in $A\left(f_{i}+\epsilon\right)=A(\alpha)$. Similarly we see that $b_{i+1}$ is the last number before 0 in $A\left(f_{i+1}-\epsilon\right)=A(\alpha)$.

Finally, we show that any beautiful arrangement of $[0, n]$ is cyclic by induction on $n$. For $n \leqslant 2$ the result is clear. Now assume that all beautiful arrangements of $[0, n-1]$ are cyclic, and consider a beautiful arrangement $A$ of $[0, n]$. The subarrangement $A_{n-1}=A \backslash\{n\}$ of $[0, n-1]$ obtained by deleting $n$ is cyclic; say $A_{n-1}=A_{n-1}(\alpha)$.

Let $\alpha$ be between the consecutive fractions $\frac{p_{1}}{q_{1}}<\frac{p_{2}}{q_{2}}$ among the irreducible fractions of denominator at most $n-1$. There is at most one fraction $\frac{i}{n}$ in $\left(\frac{p_{1}}{q_{1}}, \frac{p_{2}}{q_{2}}\right)$, since $\frac{i}{n}<\frac{i}{n-1} \leqslant \frac{i+1}{n}$ for $0<i \leqslant n-1$.

Case 1. There is no fraction with denominator $n$ between $\frac{p_{1}}{q_{1}}$ and $\frac{p_{2}}{q_{2}}$.
In this case the only cyclic arrangement extending $A_{n-1}(\alpha)$ is $A_{n}(\alpha)$. We know that $A$ and $A_{n}(\alpha)$ can only differ in the position of $n$. Assume $n$ is immediately after $x$ and before $y$ in $A_{n}(\alpha)$. Since the neighbors of 0 are $q_{1}$ and $q_{2}$ by (2), we have $x, y \geqslant 1$.


Figure 4
In $A_{n}(\alpha)$ the chord from $n-1$ to $x$ is parallel and adjacent to the chord from $n$ to $x-1$, so $n-1$ is between $x-1$ and $x$ in clockwise order, as shown in Figure 4. Similarly, $n-1$ is between $y$ and $y-1$. Therefore $x, y, x-1, n-1$, and $y-1$ occur in this order in $A_{n}(\alpha)$ and hence in $A$ (possibly with $y=x-1$ or $x=y-1$ ).

Now, $A$ may only differ from $A_{n}(\alpha)$ in the location of $n$. In $A$, since the chord from $n-1$ to $x$ and the chord from $n$ to $x-1$ do not intersect, $n$ is between $x$ and $n-1$. Similarly, $n$ is between $n-1$ and $y$. Then $n$ must be between $x$ and $y$ and $A=A_{n}(\alpha)$. Therefore $A$ is cyclic as desired.

Case 2. There is exactly one $i$ with $\frac{p_{1}}{q_{1}}<\frac{i}{n}<\frac{p_{2}}{q_{2}}$.
In this case there exist two cyclic arrangements $A_{n}\left(\alpha_{1}\right)$ and $A_{n}\left(\alpha_{2}\right)$ of the numbers $0, \ldots, n$ extending $A_{n-1}(\alpha)$, where $\frac{p_{1}}{q_{1}}<\alpha_{1}<\frac{i}{n}$ and $\frac{i}{n}<\alpha_{2}<\frac{p_{2}}{q_{2}}$. In $A_{n-1}(\alpha), 0$ is the only number between $q_{2}$ and $q_{1}$ by (2). For the same reason, $n$ is between $q_{2}$ and 0 in $A_{n}\left(\alpha_{1}\right)$, and between 0 and $q_{1}$ in $A_{n}\left(\alpha_{2}\right)$.

Letting $x=q_{2}$ and $y=q_{1}$, the argument of Case 1 tells us that $n$ must be between $x$ and $y$ in $A$. Therefore $A$ must equal $A_{n}\left(\alpha_{1}\right)$ or $A_{n}\left(\alpha_{2}\right)$, and therefore it is cyclic.

This concludes the proof that every beautiful arrangement is cyclic. It follows that there are exactly $N+1$ beautiful arrangements of $[0, n]$ as we wished to show.

C8. Players $A$ and $B$ play a paintful game on the real line. Player $A$ has a pot of paint with four units of black ink. A quantity $p$ of this ink suffices to blacken a (closed) real interval of length $p$. In every round, player $A$ picks some positive integer $m$ and provides $1 / 2^{m}$ units of ink from the pot. Player $B$ then picks an integer $k$ and blackens the interval from $k / 2^{m}$ to $(k+1) / 2^{m}$ (some parts of this interval may have been blackened before). The goal of player $A$ is to reach a situation where the pot is empty and the interval $[0,1]$ is not completely blackened.

Decide whether there exists a strategy for player $A$ to win in a finite number of moves.
(Austria)
Answer. No. Such a strategy for player $A$ does not exist.
Solution. We will present a strategy for player $B$ that guarantees that the interval $[0,1]$ is completely blackened, once the paint pot has become empty.

At the beginning of round $r$, let $x_{r}$ denote the largest real number for which the interval between 0 and $x_{r}$ has already been blackened; for completeness we define $x_{1}=0$. Let $m$ be the integer picked by player $A$ in this round; we define an integer $y_{r}$ by

$$
\frac{y_{r}}{2^{m}} \leqslant x_{r}<\frac{y_{r}+1}{2^{m}}
$$

Note that $I_{0}^{r}=\left[y_{r} / 2^{m},\left(y_{r}+1\right) / 2^{m}\right]$ is the leftmost interval that may be painted in round $r$ and that still contains some uncolored point.

Player $B$ now looks at the next interval $I_{1}^{r}=\left[\left(y_{r}+1\right) / 2^{m},\left(y_{r}+2\right) / 2^{m}\right]$. If $I_{1}^{r}$ still contains an uncolored point, then player $B$ blackens the interval $I_{1}^{r}$; otherwise he blackens the interval $I_{0}^{r}$. We make the convention that, at the beginning of the game, the interval [1,2] is already blackened; thus, if $y_{r}+1=2^{m}$, then $B$ blackens $I_{0}^{r}$.

Our aim is to estimate the amount of ink used after each round. Firstly, we will prove by induction that, if before $r$ th round the segment $[0,1]$ is not completely colored, then, before this move,
(i) the amount of ink used for the segment $\left[0, x_{r}\right]$ is at most $3 x_{r}$; and
(ii) for every $m, B$ has blackened at most one interval of length $1 / 2^{m}$ to the right of $x_{r}$.

Obviously, these conditions are satisfied for $r=0$. Now assume that they were satisfied before the $r$ th move, and consider the situation after this move; let $m$ be the number $A$ has picked at this move.

If $B$ has blackened the interval $I_{1}^{r}$ at this move, then $x_{r+1}=x_{r}$, and $(i)$ holds by the induction hypothesis. Next, had $B$ blackened before the $r$ th move any interval of length $1 / 2^{m}$ to the right of $x_{r}$, this interval would necessarily coincide with $I_{1}^{r}$. By our strategy, this cannot happen. So, condition (ii) also remains valid.

Assume now that $B$ has blackened the interval $I_{0}^{r}$ at the $r$ th move, but the interval $[0,1]$ still contains uncolored parts (which means that $I_{1}^{r}$ is contained in $[0,1]$ ). Then condition (ii) clearly remains true, and we need to check $(i)$ only. In our case, the intervals $I_{0}^{r}$ and $I_{1}^{r}$ are completely colored after the $r$ th move, so $x_{r+1}$ either reaches the right endpoint of $I_{1}$ or moves even further to the right. So, $x_{r+1}=x_{r}+\alpha$ for some $\alpha>1 / 2^{m}$.

Next, any interval blackened by $B$ before the $r$ th move which intersects $\left(x_{r}, x_{r+1}\right)$ should be contained in $\left[x_{r}, x_{r+1}\right]$; by (ii), all such intervals have different lengths not exceeding $1 / 2^{m}$, so the total amount of ink used for them is less than $2 / 2^{m}$. Thus, the amount of ink used for the segment $\left[0, x_{r+1}\right]$ does not exceed the sum of $2 / 2^{m}, 3 x_{r}$ (used for $\left[0, x_{r}\right]$ ), and $1 / 2^{m}$ used for the
segment $I_{0}^{r}$. In total it gives at most $3\left(x_{r}+1 / 2^{m}\right)<3\left(x_{r}+\alpha\right)=3 x_{r+1}$. Thus condition $(i)$ is also verified in this case. The claim is proved.

Finally, we can perform the desired estimation. Consider any situation in the game, say after the $(r-1)$ st move; assume that the segment $[0,1]$ is not completely black. By $(i i)$, in the segment $\left[x_{r}, 1\right]$ player $B$ has colored several segments of different lengths; all these lengths are negative powers of 2 not exceeding $1-x_{r}$; thus the total amount of ink used for this interval is at most $2\left(1-x_{r}\right)$. Using ( $i$, we obtain that the total amount of ink used is at most $3 x_{r}+2\left(1-x_{r}\right)<3$. Thus the pot is not empty, and therefore $A$ never wins.

Comment 1. Notice that this strategy works even if the pot contains initially only 3 units of ink.
Comment 2. There exist other strategies for $B$ allowing him to prevent emptying the pot before the whole interval is colored. On the other hand, let us mention some idea which does not work.

Player $B$ could try a strategy in which the set of blackened points in each round is an interval of the type $[0, x]$. Such a strategy cannot work (even if there is more ink available). Indeed, under the assumption that $B$ uses such a strategy, let us prove by induction on $s$ the following statement:

For any positive integer $s$, player $A$ has a strategy picking only positive integers $m \leqslant s$ in which, if player $B$ ever paints a point $x \geqslant 1-1 / 2^{s}$ then after some move, exactly the interval $\left[0,1-1 / 2^{s}\right]$ is blackened, and the amount of ink used up to this moment is at least s/2.

For the base case $s=1$, player $A$ just picks $m=1$ in the first round. If for some positive integer $k$ player $A$ has such a strategy, for $s+1$ he can first rescale his strategy to the interval [ $0,1 / 2$ ] (sending in each round half of the amount of ink he would give by the original strategy). Thus, after some round, the interval $\left[0,1 / 2-1 / 2^{s+1}\right]$ becomes blackened, and the amount of ink used is at least $s / 4$. Now player $A$ picks $m=1 / 2$, and player $B$ spends $1 / 2$ unit of ink to blacken the interval $[0,1 / 2]$. After that, player $A$ again rescales his strategy to the interval $[1 / 2,1]$, and player $B$ spends at least $s / 4$ units of ink to blacken the interval $\left[1 / 2,1-1 / 2^{s+1}\right]$, so he spends in total at least $s / 4+1 / 2+s / 4=(s+1) / 2$ units of ink.

Comment 3. In order to avoid finiteness issues, the statement could be replaced by the following one:
Players $A$ and $B$ play a paintful game on the real numbers. Player $A$ has a paint pot with four units of black ink. A quantity $p$ of this ink suffices to blacken a (closed) real interval of length $p$. In the beginning of the game, player $A$ chooses (and announces) a positive integer $N$. In every round, player $A$ picks some positive integer $m \leqslant N$ and provides $1 / 2^{m}$ units of ink from the pot. The player $B$ picks an integer $k$ and blackens the interval from $k / 2^{m}$ to $(k+1) / 2^{m}$ (some parts of this interval may happen to be blackened before). The goal of player $A$ is to reach a situation where the pot is empty and the interval $[0,1]$ is not completely blackened.
Decide whether there exists a strategy for player $A$ to win.
However, the Problem Selection Committee believes that this version may turn out to be harder than the original one.

## Geometry

G1. Let $A B C$ be an acute-angled triangle with orthocenter $H$, and let $W$ be a point on side $B C$. Denote by $M$ and $N$ the feet of the altitudes from $B$ and $C$, respectively. Denote by $\omega_{1}$ the circumcircle of $B W N$, and let $X$ be the point on $\omega_{1}$ which is diametrically opposite to $W$. Analogously, denote by $\omega_{2}$ the circumcircle of $C W M$, and let $Y$ be the point on $\omega_{2}$ which is diametrically opposite to $W$. Prove that $X, Y$ and $H$ are collinear.
(Thaliand)
Solution. Let $L$ be the foot of the altitude from $A$, and let $Z$ be the second intersection point of circles $\omega_{1}$ and $\omega_{2}$, other than $W$. We show that $X, Y, Z$ and $H$ lie on the same line.

Due to $\angle B N C=\angle B M C=90^{\circ}$, the points $B, C, N$ and $M$ are concyclic; denote their circle by $\omega_{3}$. Observe that the line $W Z$ is the radical axis of $\omega_{1}$ and $\omega_{2}$; similarly, $B N$ is the radical axis of $\omega_{1}$ and $\omega_{3}$, and $C M$ is the radical axis of $\omega_{2}$ and $\omega_{3}$. Hence $A=B N \cap C M$ is the radical center of the three circles, and therefore $W Z$ passes through $A$.

Since $W X$ and $W Y$ are diameters in $\omega_{1}$ and $\omega_{2}$, respectively, we have $\angle W Z X=\angle W Z Y=90^{\circ}$, so the points $X$ and $Y$ lie on the line through $Z$, perpendicular to $W Z$.


The quadrilateral $B L H N$ is cyclic, because it has two opposite right angles. From the power of $A$ with respect to the circles $\omega_{1}$ and $B L H N$ we find $A L \cdot A H=A B \cdot A N=A W \cdot A Z$. If $H$ lies on the line $A W$ then this implies $H=Z$ immediately. Otherwise, by $\frac{A Z}{A H}=\frac{A L}{A W}$ the triangles $A H Z$ and $A W L$ are similar. Then $\angle H Z A=\angle W L A=90^{\circ}$, so the point $H$ also lies on the line $X Y Z$.

Comment. The original proposal also included a second statement:
Let $P$ be the point on $\omega_{1}$ such that $W P$ is parallel to $C N$, and let $Q$ be the point on $\omega_{2}$ such that $W Q$ is parallel to $B M$. Prove that $P, Q$ and $H$ are collinear if and only if $B W=C W$ or $A W \perp B C$.

The Problem Selection Committee considered the first part more suitable for the competition.

G2. Let $\omega$ be the circumcircle of a triangle $A B C$. Denote by $M$ and $N$ the midpoints of the sides $A B$ and $A C$, respectively, and denote by $T$ the midpoint of the arc $B C$ of $\omega$ not containing $A$. The circumcircles of the triangles $A M T$ and $A N T$ intersect the perpendicular bisectors of $A C$ and $A B$ at points $X$ and $Y$, respectively; assume that $X$ and $Y$ lie inside the triangle $A B C$. The lines $M N$ and $X Y$ intersect at $K$. Prove that $K A=K T$.
(Iran)
Solution 1. Let $O$ be the center of $\omega$, thus $O=M Y \cap N X$. Let $\ell$ be the perpendicular bisector of $A T$ (it also passes through $O$ ). Denote by $r$ the operation of reflection about $\ell$. Since $A T$ is the angle bisector of $\angle B A C$, the line $r(A B)$ is parallel to $A C$. Since $O M \perp A B$ and $O N \perp A C$, this means that the line $r(O M)$ is parallel to the line $O N$ and passes through $O$, so $r(O M)=O N$. Finally, the circumcircle $\gamma$ of the triangle $A M T$ is symmetric about $\ell$, so $r(\gamma)=\gamma$. Thus the point $M$ maps to the common point of $O N$ with the arc $A M T$ of $\gamma-$ that is, $r(M)=X$.

Similarly, $r(N)=Y$. Thus, we get $r(M N)=X Y$, and the common point $K$ of $M N$ nd $X Y$ lies on $\ell$. This means exactly that $K A=K T$.


Solution 2. Let $L$ be the second common point of the line $A C$ with the circumcircle $\gamma$ of the triangle $A M T$. From the cyclic quadrilaterals $A B T C$ and $A M T L$ we get $\angle B T C=180^{\circ}-$ $\angle B A C=\angle M T L$, which implies $\angle B T M=\angle C T L$. Since $A T$ is an angle bisector in these quadrilaterals, we have $B T=T C$ and $M T=T L$. Thus the triangles $B T M$ and $C T L$ are congruent, so $C L=B M=A M$.

Let $X^{\prime}$ be the common point of the line $N X$ with the external bisector of $\angle B A C$; notice that it lies outside the triangle $A B C$. Then we have $\angle T A X^{\prime}=90^{\circ}$ and $X^{\prime} A=X^{\prime} C$, so we get $\angle X^{\prime} A M=90^{\circ}+\angle B A C / 2=180^{\circ}-\angle X^{\prime} A C=180^{\circ}-\angle X^{\prime} C A=\angle X^{\prime} C L$. Thus the triangles $X^{\prime} A M$ and $X^{\prime} C L$ are congruent, and therefore

$$
\angle M X^{\prime} L=\angle A X^{\prime} C+\left(\angle C X^{\prime} L-\angle A X^{\prime} M\right)=\angle A X^{\prime} C=180^{\circ}-2 \angle X^{\prime} A C=\angle B A C=\angle M A L .
$$

This means that $X^{\prime}$ lies on $\gamma$.
Thus we have $\angle T X N=\angle T X X^{\prime}=\angle T A X^{\prime}=90^{\circ}$, so $T X \| A C$. Then $\angle X T A=\angle T A C=$ $\angle T A M$, so the cyclic quadrilateral $M A T X$ is an isosceles trapezoid. Similarly, $N A T Y$ is an isosceles trapezoid, so again the lines $M N$ and $X Y$ are the reflections of each other about the perpendicular bisector of $A T$. Thus $K$ belongs to this perpendicular bisector.


Comment. There are several different ways of showing that the points $X$ and $M$ are symmetrical with respect to $\ell$. For instance, one can show that the quadrilaterals $A M O N$ and $T X O Y$ are congruent. We chose Solution 1 as a simple way of doing it. On the other hand, Solution 2 shows some other interesting properties of the configuration.

Let us define $Y^{\prime}$, analogously to $X^{\prime}$, as the common point of $M Y$ and the external bisector of $\angle B A C$. One may easily see that in general the lines $M N$ and $X^{\prime} Y^{\prime}$ (which is the external bisector of $\angle B A C$ ) do not intersect on the perpendicular bisector of $A T$. Thus, any solution should involve some argument using the choice of the intersection points $X$ and $Y$.

G3. In a triangle $A B C$, let $D$ and $E$ be the feet of the angle bisectors of angles $A$ and $B$, respectively. A rhombus is inscribed into the quadrilateral $A E D B$ (all vertices of the rhombus lie on different sides of $A E D B$ ). Let $\varphi$ be the non-obtuse angle of the rhombus. Prove that $\varphi \leqslant \max \{\angle B A C, \angle A B C\}$.
(Serbia)
Solution 1. Let $K, L, M$, and $N$ be the vertices of the rhombus lying on the sides $A E, E D, D B$, and $B A$, respectively. Denote by $d(X, Y Z)$ the distance from a point $X$ to a line $Y Z$. Since $D$ and $E$ are the feet of the bisectors, we have $d(D, A B)=d(D, A C), d(E, A B)=d(E, B C)$, and $d(D, B C)=d(E, A C)=0$, which implies

$$
d(D, A C)+d(D, B C)=d(D, A B) \quad \text { and } \quad d(E, A C)+d(E, B C)=d(E, A B)
$$

Since $L$ lies on the segment $D E$ and the relation $d(X, A C)+d(X, B C)=d(X, A B)$ is linear in $X$ inside the triangle, these two relations imply

$$
\begin{equation*}
d(L, A C)+d(L, B C)=d(L, A B) \tag{1}
\end{equation*}
$$

Denote the angles as in the figure below, and denote $a=K L$. Then we have $d(L, A C)=a \sin \mu$ and $d(L, B C)=a \sin \nu$. Since $K L M N$ is a parallelogram lying on one side of $A B$, we get

$$
d(L, A B)=d(L, A B)+d(N, A B)=d(K, A B)+d(M, A B)=a(\sin \delta+\sin \varepsilon)
$$

Thus the condition (1) reads

$$
\begin{equation*}
\sin \mu+\sin \nu=\sin \delta+\sin \varepsilon \tag{2}
\end{equation*}
$$



If one of the angles $\alpha$ and $\beta$ is non-acute, then the desired inequality is trivial. So we assume that $\alpha, \beta<\pi / 2$. It suffices to show then that $\psi=\angle N K L \leqslant \max \{\alpha, \beta\}$.

Assume, to the contrary, that $\psi>\max \{\alpha, \beta\}$. Since $\mu+\psi=\angle C K N=\alpha+\delta$, by our assumption we obtain $\mu=(\alpha-\psi)+\delta<\delta$. Similarly, $\nu<\varepsilon$. Next, since $K N \| M L$, we have $\beta=\delta+\nu$, so $\delta<\beta<\pi / 2$. Similarly, $\varepsilon<\pi / 2$. Finally, by $\mu<\delta<\pi / 2$ and $\nu<\varepsilon<\pi / 2$, we obtain

$$
\sin \mu<\sin \delta \quad \text { and } \quad \sin \nu<\sin \varepsilon
$$

This contradicts (2).
Comment. One can see that the equality is achieved if $\alpha=\beta$ for every rhombus inscribed into the quadrilateral $A E D B$.

G4. Let $A B C$ be a triangle with $\angle B>\angle C$. Let $P$ and $Q$ be two different points on line $A C$ such that $\angle P B A=\angle Q B A=\angle A C B$ and $A$ is located between $P$ and $C$. Suppose that there exists an interior point $D$ of segment $B Q$ for which $P D=P B$. Let the ray $A D$ intersect the circle $A B C$ at $R \neq A$. Prove that $Q B=Q R$.
(Georgia)
Solution 1. Denote by $\omega$ the circumcircle of the triangle $A B C$, and let $\angle A C B=\gamma$. Note that the condition $\gamma<\angle C B A$ implies $\gamma<90^{\circ}$. Since $\angle P B A=\gamma$, the line $P B$ is tangent to $\omega$, so $P A \cdot P C=P B^{2}=P D^{2}$. By $\frac{P A}{P D}=\frac{P D}{P C}$ the triangles $P A D$ and $P D C$ are similar, and $\angle A D P=\angle D C P$.

Next, since $\angle A B Q=\angle A C B$, the triangles $A B C$ and $A Q B$ are also similar. Then $\angle A Q B=$ $\angle A B C=\angle A R C$, which means that the points $D, R, C$, and $Q$ are concyclic. Therefore $\angle D R Q=$ $\angle D C Q=\angle A D P$.


Figure 1
Now from $\angle A R B=\angle A C B=\gamma$ and $\angle P D B=\angle P B D=2 \gamma$ we get

$$
\angle Q B R=\angle A D B-\angle A R B=\angle A D P+\angle P D B-\angle A R B=\angle D R Q+\gamma=\angle Q R B
$$

so the triangle $Q R B$ is isosceles, which yields $Q B=Q R$.
Solution 2. Again, denote by $\omega$ the circumcircle of the triangle $A B C$. Denote $\angle A C B=\gamma$. Since $\angle P B A=\gamma$, the line $P B$ is tangent to $\omega$.

Let $E$ be the second intersection point of $B Q$ with $\omega$. If $V^{\prime}$ is any point on the ray $C E$ beyond $E$, then $\angle B E V^{\prime}=180^{\circ}-\angle B E C=180^{\circ}-\angle B A C=\angle P A B$; together with $\angle A B Q=$ $\angle P B A$ this shows firstly, that the rays $B A$ and $C E$ intersect at some point $V$, and secondly that the triangle $V E B$ is similar to the triangle $P A B$. Thus we have $\angle B V E=\angle B P A$. Next, $\angle A E V=\angle B E V-\gamma=\angle P A B-\angle A B Q=\angle A Q B$; so the triangles $P B Q$ and $V A E$ are also similar.

Let $P H$ be an altitude in the isosceles triangle $P B D$; then $B H=H D$. Let $G$ be the intersection point of $P H$ and $A B$. By the symmetry with respect to $P H$, we have $\angle B D G=\angle D B G=\gamma=$ $\angle B E A$; thus $D G \| A E$ and hence $\frac{B G}{G A}=\frac{B D}{D E}$. Thus the points $G$ and $D$ correspond to each other in the similar triangles $P A B$ and $V E B$, so $\angle D V B=\angle G P B=90^{\circ}-\angle P B Q=90^{\circ}-\angle V A E$. Thus $V D \perp A E$.

Let $T$ be the common point of $V D$ and $A E$, and let $D S$ be an altitude in the triangle $B D R$. The points $S$ and $T$ are the feet of corresponding altitudes in the similar triangles $A D E$ and $B D R$, so $\frac{B S}{S R}=\frac{A T}{T E}$. On the other hand, the points $T$ and $H$ are feet of corresponding altitudes in the similar triangles $V A E$ and $P B Q$, so $\frac{A T}{T E}=\frac{B H}{H Q}$. Thus $\frac{B S}{S R}=\frac{A T}{T E}=\frac{B H}{H Q}$, and the triangles $B H S$ and $B Q R$ are similar.

Finally, $S H$ is a median in the right-angled triangle $S B D$; so $B H=H S$, and hence $B Q=Q R$.


Figure 2

Solution 3. Denote by $\omega$ and $O$ the circumcircle of the triangle $A B C$ and its center, respectively. From the condition $\angle P B A=\angle B C A$ we know that $B P$ is tangent to $\omega$.

Let $E$ be the second point of intersection of $\omega$ and $B D$. Due to the isosceles triangle $B D P$, the tangent of $\omega$ at $E$ is parallel to $D P$ and consequently it intersects $B P$ at some point $L$. Of course, $P D \| L E$. Let $M$ be the midpoint of $B E$, and let $H$ be the midpoint of $B R$. Notice that $\angle A E B=\angle A C B=\angle A B Q=\angle A B E$, so $A$ lies on the perpendicular bisector of $B E$; thus the points $L, A, M$, and $O$ are collinear. Let $\omega_{1}$ be the circle with diameter $B O$. Let $Q^{\prime}=H O \cap B E$; since $H O$ is the perpendicular bisector of $B R$, the statement of the problem is equivalent to $Q^{\prime}=Q$.

Consider the following sequence of projections (see Fig. 3).

1. Project the line $B E$ to the line $L B$ through the center $A$. (This maps $Q$ to $P$.)
2. Project the line $L B$ to $B E$ in parallel direction with $L E .(P \mapsto D$.)
3. Project the line $B E$ to the circle $\omega$ through its point $A .(D \mapsto R$.)
4. Scale $\omega$ by the ratio $\frac{1}{2}$ from the point $B$ to the circle $\omega_{1}$. $(R \mapsto H$.
5. Project $\omega_{1}$ to the line $B E$ through its point $O$. $\left(H \mapsto Q^{\prime}\right.$.)

We prove that the composition of these transforms, which maps the line $B E$ to itself, is the identity. To achieve this, it suffices to show three fixed points. An obvious fixed point is $B$ which is fixed by all the transformations above. Another fixed point is $M$, its path being $M \mapsto L \mapsto$ $E \mapsto E \mapsto M \mapsto M$.


Figure 3


Figure 4

In order to show a third fixed point, draw a line parallel with $L E$ through $A$; let that line intersect $B E, L B$ and $\omega$ at $X, Y$ and $Z \neq A$, respectively (see Fig. 4). We show that $X$ is a fixed point. The images of $X$ at the first three transformations are $X \mapsto Y \mapsto X \mapsto Z$. From $\angle X B Z=\angle E A Z=\angle A E L=\angle L B A=\angle B Z X$ we can see that the triangle $X B Z$ is isosceles. Let $U$ be the midpoint of $B Z$; then the last two transformations do $Z \mapsto U \mapsto X$, and the point $X$ is fixed.

Comment. Verifying that the point $E$ is fixed seems more natural at first, but it appears to be less straightforward. Here we outline a possible proof.

Let the images of $E$ at the first three transforms above be $F, G$ and $I$. After comparing the angles depicted in Fig. 5 (noticing that the quadrilateral $A F B G$ is cyclic) we can observe that the tangent $L E$ of $\omega$ is parallel to $B I$. Then, similarly to the above reasons, the point $E$ is also fixed.


Figure 5

G5. Let $A B C D E F$ be a convex hexagon with $A B=D E, B C=E F, C D=F A$, and $\angle A-\angle D=\angle C-\angle F=\angle E-\angle B$. Prove that the diagonals $A D, B E$, and $C F$ are concurrent.
(Ukraine)
In all three solutions, we denote $\theta=\angle A-\angle D=\angle C-\angle F=\angle E-\angle B$ and assume without loss of generality that $\theta \geqslant 0$.
Solution 1. Let $x=A B=D E, y=C D=F A, z=E F=B C$. Consider the points $P, Q$, and $R$ such that the quadrilaterals $C D E P, E F A Q$, and $A B C R$ are parallelograms. We compute

$$
\begin{aligned}
\angle P E Q & =\angle F E Q+\angle D E P-\angle E=\left(180^{\circ}-\angle F\right)+\left(180^{\circ}-\angle D\right)-\angle E \\
& =360^{\circ}-\angle D-\angle E-\angle F=\frac{1}{2}(\angle A+\angle B+\angle C-\angle D-\angle E-\angle F)=\theta / 2
\end{aligned}
$$

Similarly, $\angle Q A R=\angle R C P=\theta / 2$.


If $\theta=0$, since $\triangle R C P$ is isosceles, $R=P$. Therefore $A B\|R C=P C\| E D$, so $A B D E$ is a parallelogram. Similarly, $B C E F$ and $C D F A$ are parallelograms. It follows that $A D, B E$ and $C F$ meet at their common midpoint.

Now assume $\theta>0$. Since $\triangle P E Q, \triangle Q A R$, and $\triangle R C P$ are isosceles and have the same angle at the apex, we have $\triangle P E Q \sim \triangle Q A R \sim \triangle R C P$ with ratios of similarity $y: z: x$. Thus

$$
\begin{equation*}
\triangle P Q R \text { is similar to the triangle with sidelengths } y, z, \text { and } x . \tag{1}
\end{equation*}
$$

Next, notice that

$$
\frac{R Q}{Q P}=\frac{z}{y}=\frac{R A}{A F}
$$

and, using directed angles between rays,

$$
\begin{aligned}
\Varangle(R Q, Q P) & =\Varangle(R Q, Q E)+\not(Q E, Q P) \\
& =\Varangle(R Q, Q E)+\not(R A, R Q)=\Varangle(R A, Q E)=\Varangle(R A, A F) .
\end{aligned}
$$

Thus $\triangle P Q R \sim \triangle F A R$. Since $F A=y$ and $A R=z$, (1) then implies that $F R=x$. Similarly $F P=x$. Therefore $C R F P$ is a rhombus.

We conclude that $C F$ is the perpendicular bisector of $P R$. Similarly, $B E$ is the perpendicular bisector of $P Q$ and $A D$ is the perpendicular bisector of $Q R$. It follows that $A D, B E$, and $C F$ are concurrent at the circumcenter of $P Q R$.

Solution 2. Let $X=C D \cap E F, Y=E F \cap A B, Z=A B \cap C D, X^{\prime}=F A \cap B C, Y^{\prime}=$ $B C \cap D E$, and $Z^{\prime}=D E \cap F A$. From $\angle A+\angle B+\angle C=360^{\circ}+\theta / 2$ we get $\angle A+\angle B>180^{\circ}$ and $\angle B+\angle C>180^{\circ}$, so $Z$ and $X^{\prime}$ are respectively on the opposite sides of $B C$ and $A B$ from the hexagon. Similar conclusions hold for $X, Y, Y^{\prime}$, and $Z^{\prime}$. Then

$$
\angle Y Z X=\angle B+\angle C-180^{\circ}=\angle E+\angle F-180^{\circ}=\angle Y^{\prime} Z^{\prime} X^{\prime}
$$

and similarly $\angle Z X Y=\angle Z^{\prime} X^{\prime} Y^{\prime}$ and $\angle X Y Z=\angle X^{\prime} Y^{\prime} Z^{\prime}$, so $\triangle X Y Z \sim \triangle X^{\prime} Y^{\prime} Z^{\prime}$. Thus there is a rotation $R$ which sends $\triangle X Y Z$ to a triangle with sides parallel to $\triangle X^{\prime} Y^{\prime} Z^{\prime}$. Since $A B=D E$ we have $R(\overrightarrow{A B})=\overrightarrow{D E}$. Similarly, $R(\overrightarrow{C D})=\overrightarrow{F A}$ and $R(\overrightarrow{E F})=\overrightarrow{B C}$. Therefore

$$
\overrightarrow{0}=\overrightarrow{A B}+\overrightarrow{B C}+\overrightarrow{C D}+\overrightarrow{D E}+\overrightarrow{E F}+\overrightarrow{F A}=(\overrightarrow{A B}+\overrightarrow{C D}+\overrightarrow{E F})+R(\overrightarrow{A B}+\overrightarrow{C D}+\overrightarrow{E F})
$$

If $R$ is a rotation by $180^{\circ}$, then any two opposite sides of our hexagon are equal and parallel, so the three diagonals meet at their common midpoint. Otherwise, we must have

$$
\overrightarrow{A B}+\overrightarrow{C D}+\overrightarrow{E F}=\overrightarrow{0}
$$

or else we would have two vectors with different directions whose sum is $\overrightarrow{0}$.


This allows us to consider a triangle $L M N$ with $\overrightarrow{L M}=\overrightarrow{E F}, \overrightarrow{M N}=\overrightarrow{A B}$, and $\overrightarrow{N L}=\overrightarrow{C D}$. Let $O$ be the circumcenter of $\triangle L M N$ and consider the points $O_{1}, O_{2}, O_{3}$ such that $\triangle A O_{1} B, \triangle C O_{2} D$, and $\triangle E O_{3} F$ are translations of $\triangle M O N, \triangle N O L$, and $\triangle L O M$, respectively. Since $F O_{3}$ and $A O_{1}$ are translations of $M O$, quadrilateral $A F O_{3} O_{1}$ is a parallelogram and $O_{3} O_{1}=F A=C D=N L$. Similarly, $O_{1} O_{2}=L M$ and $O_{2} O_{3}=M N$. Therefore $\triangle O_{1} O_{2} O_{3} \cong \triangle L M N$. Moreover, by means of the rotation $R$ one may check that these triangles have the same orientation.

Let $T$ be the circumcenter of $\triangle O_{1} O_{2} O_{3}$. We claim that $A D, B E$, and $C F$ meet at $T$. Let us show that $C, T$, and $F$ are collinear. Notice that $C O_{2}=O_{2} T=T O_{3}=O_{3} F$ since they are all equal to the circumradius of $\triangle L M N$. Therefore $\triangle T O_{3} F$ and $\triangle C O_{2} T$ are isosceles. Using directed angles between rays again, we get

$$
\begin{equation*}
\Varangle\left(T F, T O_{3}\right)=\Varangle\left(F O_{3}, F T\right) \quad \text { and } \quad \nsucceq\left(T O_{2}, T C\right)=\nsucceq\left(C T, C O_{2}\right) . \tag{2}
\end{equation*}
$$

Also, $T$ and $O$ are the circumcenters of the congruent triangles $\triangle O_{1} O_{2} O_{3}$ and $\triangle L M N$ so we have $\Varangle\left(T O_{3}, T O_{2}\right)=\Varangle(O N, O M)$. Since $\mathrm{CO}_{2}$ and $\mathrm{FO}_{3}$ are translations of $N O$ and $M O$ respectively, this implies

$$
\begin{equation*}
\Varangle\left(T O_{3}, T O_{2}\right)=\Varangle\left(C O_{2}, F O_{3}\right) . \tag{3}
\end{equation*}
$$

Adding the three equations in (2) and (3) gives

$$
\npreceq(T F, T C)=\npreceq(C T, F T)=-\nless(T F, T C)
$$

which implies that $T$ is on $C F$. Analogous arguments show that it is on $A D$ and $B E$ also. The desired result follows.

Solution 3. Place the hexagon on the complex plane, with $A$ at the origin and vertices labelled clockwise. Now $A, B, C, D, E, F$ represent the corresponding complex numbers. Also consider the complex numbers $a, b, c, a^{\prime}, b^{\prime}, c^{\prime}$ given by $B-A=a, D-C=b, F-E=c, E-D=a^{\prime}$, $A-F=b^{\prime}$, and $C-B=c^{\prime}$. Let $k=|a| /|b|$. From $a / b^{\prime}=-k e^{i \angle A}$ and $a^{\prime} / b=-k e^{i \angle D}$ we get that $\left(a^{\prime} / a\right)\left(b^{\prime} / b\right)=e^{-i \theta}$ and similarly $\left(b^{\prime} / b\right)\left(c^{\prime} / c\right)=e^{-i \theta}$ and $\left(c^{\prime} / c\right)\left(a^{\prime} / a\right)=e^{-i \theta}$. It follows that $a^{\prime}=a r$, $b^{\prime}=b r$, and $c^{\prime}=c r$ for a complex number $r$ with $|r|=1$, as shown below.


We have

$$
0=a+c r+b+a r+c+b r=(a+b+c)(1+r)
$$

If $r=-1$, then the hexagon is centrally symmetric and its diagonals intersect at its center of symmetry. Otherwise

$$
a+b+c=0
$$

Therefore

$$
A=0, \quad B=a, \quad C=a+c r, \quad D=c(r-1), \quad E=-b r-c, \quad F=-b r .
$$

Now consider a point $W$ on $A D$ given by the complex number $c(r-1) \lambda$, where $\lambda$ is a real number with $0<\lambda<1$. Since $D \neq A$, we have $r \neq 1$, so we can define $s=1 /(r-1)$. From $r \bar{r}=|r|^{2}=1$ we get

$$
1+s=\frac{r}{r-1}=\frac{r}{r-r \bar{r}}=\frac{1}{1-\bar{r}}=-\bar{s} .
$$

Now,

$$
\begin{aligned}
W \text { is on } B E & \Longleftrightarrow c(r-1) \lambda-a\|a-(-b r-c)=b(r-1) \Longleftrightarrow c \lambda-a s\| b \\
& \Longleftrightarrow-a \lambda-b \lambda-a s\|b \Longleftrightarrow a(\lambda+s)\| b .
\end{aligned}
$$

One easily checks that $r \neq \pm 1$ implies that $\lambda+s \neq 0$ since $s$ is not real. On the other hand,

$$
\begin{aligned}
W \text { on } C F & \Longleftrightarrow c(r-1) \lambda+b r\|-b r-(a+c r)=a(r-1) \Longleftrightarrow c \lambda+b(1+s)\| a \\
& \Longleftrightarrow-a \lambda-b \lambda-b \bar{s}\|a \Longleftrightarrow b(\lambda+\bar{s})\| a \Longleftrightarrow b \| a(\lambda+s),
\end{aligned}
$$

where in the last step we use that $(\lambda+s)(\lambda+\bar{s})=|\lambda+s|^{2} \in \mathbb{R}_{>0}$. We conclude that $A D \cap B E=$ $C F \cap B E$, and the desired result follows.

G6. Let the excircle of the triangle $A B C$ lying opposite to $A$ touch its side $B C$ at the point $A_{1}$. Define the points $B_{1}$ and $C_{1}$ analogously. Suppose that the circumcentre of the triangle $A_{1} B_{1} C_{1}$ lies on the circumcircle of the triangle $A B C$. Prove that the triangle $A B C$ is right-angled.
(Russia)
Solution 1. Denote the circumcircles of the triangles $A B C$ and $A_{1} B_{1} C_{1}$ by $\Omega$ and $\Gamma$, respectively. Denote the midpoint of the arc $C B$ of $\Omega$ containing $A$ by $A_{0}$, and define $B_{0}$ as well as $C_{0}$ analogously. By our hypothesis the centre $Q$ of $\Gamma$ lies on $\Omega$.
Lemma. One has $A_{0} B_{1}=A_{0} C_{1}$. Moreover, the points $A, A_{0}, B_{1}$, and $C_{1}$ are concyclic. Finally, the points $A$ and $A_{0}$ lie on the same side of $B_{1} C_{1}$. Similar statements hold for $B$ and $C$.
Proof. Let us consider the case $A=A_{0}$ first. Then the triangle $A B C$ is isosceles at $A$, which implies $A B_{1}=A C_{1}$ while the remaining assertions of the Lemma are obvious. So let us suppose $A \neq A_{0}$ from now on.

By the definition of $A_{0}$, we have $A_{0} B=A_{0} C$. It is also well known and easy to show that $B C_{1}=$ $C B_{1}$. Next, we have $\angle C_{1} B A_{0}=\angle A B A_{0}=\angle A C A_{0}=\angle B_{1} C A_{0}$. Hence the triangles $A_{0} B C_{1}$ and $A_{0} C B_{1}$ are congruent. This implies $A_{0} C_{1}=A_{0} B_{1}$, establishing the first part of the Lemma. It also follows that $\angle A_{0} C_{1} A=\angle A_{0} B_{1} A$, as these are exterior angles at the corresponding vertices $C_{1}$ and $B_{1}$ of the congruent triangles $A_{0} B C_{1}$ and $A_{0} C B_{1}$. For that reason the points $A, A_{0}, B_{1}$, and $C_{1}$ are indeed the vertices of some cyclic quadrilateral two opposite sides of which are $A A_{0}$ and $B_{1} C_{1}$.

Now we turn to the solution. Evidently the points $A_{1}, B_{1}$, and $C_{1}$ lie interior to some semicircle arc of $\Gamma$, so the triangle $A_{1} B_{1} C_{1}$ is obtuse-angled. Without loss of generality, we will assume that its angle at $B_{1}$ is obtuse. Thus $Q$ and $B_{1}$ lie on different sides of $A_{1} C_{1}$; obviously, the same holds for the points $B$ and $B_{1}$. So, the points $Q$ and $B$ are on the same side of $A_{1} C_{1}$.

Notice that the perpendicular bisector of $A_{1} C_{1}$ intersects $\Omega$ at two points lying on different sides of $A_{1} C_{1}$. By the first statement from the Lemma, both points $B_{0}$ and $Q$ are among these points of intersection; since they share the same side of $A_{1} C_{1}$, they coincide (see Figure 1).


Figure 1

Now, by the first part of the Lemma again, the lines $Q A_{0}$ and $Q C_{0}$ are the perpendicular bisectors of $B_{1} C_{1}$ and $A_{1} B_{1}$, respectively. Thus

$$
\angle C_{1} B_{0} A_{1}=\angle C_{1} B_{0} B_{1}+\angle B_{1} B_{0} A_{1}=2 \angle A_{0} B_{0} B_{1}+2 \angle B_{1} B_{0} C_{0}=2 \angle A_{0} B_{0} C_{0}=180^{\circ}-\angle A B C,
$$

recalling that $A_{0}$ and $C_{0}$ are the midpoints of the $\operatorname{arcs} C B$ and $B A$, respectively.
On the other hand, by the second part of the Lemma we have

$$
\angle C_{1} B_{0} A_{1}=\angle C_{1} B A_{1}=\angle A B C
$$

From the last two equalities, we get $\angle A B C=90^{\circ}$, whereby the problem is solved.
Solution 2. Let $Q$ again denote the centre of the circumcircle of the triangle $A_{1} B_{1} C_{1}$, that lies on the circumcircle $\Omega$ of the triangle $A B C$. We first consider the case where $Q$ coincides with one of the vertices of $A B C$, say $Q=B$. Then $B C_{1}=B A_{1}$ and consequently the triangle $A B C$ is isosceles at $B$. Moreover we have $B C_{1}=B_{1} C$ in any triangle, and hence $B B_{1}=B C_{1}=B_{1} C$; similarly, $B B_{1}=B_{1} A$. It follows that $B_{1}$ is the centre of $\Omega$ and that the triangle $A B C$ has a right angle at $B$.

So from now on we may suppose $Q \notin\{A, B, C\}$. We start with the following well known fact. Lemma. Let $X Y Z$ and $X^{\prime} Y^{\prime} Z^{\prime}$ be two triangles with $X Y=X^{\prime} Y^{\prime}$ and $Y Z=Y^{\prime} Z^{\prime}$.
(i) If $X Z \neq X^{\prime} Z^{\prime}$ and $\angle Y Z X=\angle Y^{\prime} Z^{\prime} X^{\prime}$, then $\angle Z X Y+\angle Z^{\prime} X^{\prime} Y^{\prime}=180^{\circ}$.
(ii) If $\angle Y Z X+\angle X^{\prime} Z^{\prime} Y^{\prime}=180^{\circ}$, then $\angle Z X Y=\angle Y^{\prime} X^{\prime} Z^{\prime}$.

Proof. For both parts, we may move the triangle $X Y Z$ through the plane until $Y=Y^{\prime}$ and $Z=Z^{\prime}$. Possibly after reflecting one of the two triangles about $Y Z$, we may also suppose that $X$ and $X^{\prime}$ lie on the same side of $Y Z$ if we are in case $(i)$ and on different sides if we are in case (ii). In both cases, the points $X, Z$, and $X^{\prime}$ are collinear due to the angle condition (see Fig. 2). Moreover we have $X \neq X^{\prime}$, because in case $(i)$ we assumed $X Z \neq X^{\prime} Z^{\prime}$ and in case (ii) these points even lie on different sides of $Y Z$. Thus the triangle $X X^{\prime} Y$ is isosceles at $Y$. The claim now follows by considering the equal angles at its base.


Figure 2( $i$ )


Figure 2(ii)

Relabeling the vertices of the triangle $A B C$ if necessary we may suppose that $Q$ lies in the interior of the arc $A B$ of $\Omega$ not containing $C$. We will sometimes use tacitly that the six triangles $Q B A_{1}, Q A_{1} C, Q C B_{1}, Q B_{1} A, Q C_{1} A$, and $Q B C_{1}$ have the same orientation.

As $Q$ cannot be the circumcentre of the triangle $A B C$, it is impossible that $Q A=Q B=Q C$ and thus we may also suppose that $Q C \neq Q B$. Now the above Lemma $(i)$ is applicable to the triangles $Q B_{1} C$ and $Q C_{1} B$, since $Q B_{1}=Q C_{1}$ and $B_{1} C=C_{1} B$, while $\angle B_{1} C Q=\angle C_{1} B Q$ holds as both angles appear over the same side of the chord $Q A$ in $\Omega$ (see Fig. 3). So we get

$$
\begin{equation*}
\angle C Q B_{1}+\angle B Q C_{1}=180^{\circ} . \tag{1}
\end{equation*}
$$

We claim that $Q C=Q A$. To see this, let us assume for the sake of a contradiction that $Q C \neq Q A$. Then arguing similarly as before but now with the triangles $Q A_{1} C$ and $Q C_{1} A$ we get

$$
\angle A_{1} Q C+\angle C_{1} Q A=180^{\circ} .
$$

Adding this equation to (1), we get $\angle A_{1} Q B_{1}+\angle B Q A=360^{\circ}$, which is absurd as both summands lie in the interval $\left(0^{\circ}, 180^{\circ}\right)$.

This proves $Q C=Q A$; so the triangles $Q A_{1} C$ and $Q C_{1} A$ are congruent their sides being equal, which in turn yields

$$
\begin{equation*}
\angle A_{1} Q C=\angle C_{1} Q A . \tag{2}
\end{equation*}
$$

Finally our Lemma ( $i$ i $)$ is applicable to the triangles $Q A_{1} B$ and $Q B_{1} A$. Indeed we have $Q A_{1}=Q B_{1}$ and $A_{1} B=B_{1} A$ as usual, and the angle condition $\angle A_{1} B Q+\angle Q A B_{1}=180^{\circ}$ holds as $A$ and $B$ lie on different sides of the chord $Q C$ in $\Omega$. Consequently we have

$$
\begin{equation*}
\angle B Q A_{1}=\angle B_{1} Q A \tag{3}
\end{equation*}
$$

From (1) and (3) we get

$$
\left(\angle B_{1} Q C+\angle B_{1} Q A\right)+\left(\angle C_{1} Q B-\angle B Q A_{1}\right)=180^{\circ},
$$

i.e. $\angle C Q A+\angle A_{1} Q C_{1}=180^{\circ}$. In light of (2) this may be rewritten as $2 \angle C Q A=180^{\circ}$ and as $Q$ lies on $\Omega$ this implies that the triangle $A B C$ has a right angle at $B$.


Figure 3

Comment 1. One may also check that $Q$ is in the interior of $\Omega$ if and only if the triangle $A B C$ is acute-angled.

Comment 2. The original proposal asked to prove the converse statement as well: if the triangle $A B C$ is right-angled, then the point $Q$ lies on its circumcircle. The Problem Selection Committee thinks that the above simplified version is more suitable for the competition.

## Number Theory

N1. Let $\mathbb{Z}_{>0}$ be the set of positive integers. Find all functions $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ such that

$$
m^{2}+f(n) \mid m f(m)+n
$$

for all positive integers $m$ and $n$.
(Malaysia)
Answer. $f(n)=n$.
Solution 1. Setting $m=n=2$ tells us that $4+f(2) \mid 2 f(2)+2$. Since $2 f(2)+2<2(4+f(2))$, we must have $2 f(2)+2=4+f(2)$, so $f(2)=2$. Plugging in $m=2$ then tells us that $4+f(n) \mid 4+n$, which implies that $f(n) \leqslant n$ for all $n$.

Setting $m=n$ gives $n^{2}+f(n) \mid n f(n)+n$, so $n f(n)+n \geqslant n^{2}+f(n)$ which we rewrite as $(n-1)(f(n)-n) \geqslant 0$. Therefore $f(n) \geqslant n$ for all $n \geqslant 2$. This is trivially true for $n=1$ also.

It follows that $f(n)=n$ for all $n$. This function obviously satisfies the desired property.
Solution 2. Setting $m=f(n)$ we get $f(n)(f(n)+1) \mid f(n) f(f(n))+n$. This implies that $f(n) \mid n$ for all $n$.

Now let $m$ be any positive integer, and let $p>2 m^{2}$ be a prime number. Note that $p>m f(m)$ also. Plugging in $n=p-m f(m)$ we learn that $m^{2}+f(n)$ divides $p$. Since $m^{2}+f(n)$ cannot equal 1 , it must equal $p$. Therefore $p-m^{2}=f(n) \mid n=p-m f(m)$. But $p-m f(m)<p<2\left(p-m^{2}\right)$, so we must have $p-m f(m)=p-m^{2}$, i.e., $f(m)=m$.

Solution 3. Plugging $m=1$ we obtain $1+f(n) \leqslant f(1)+n$, so $f(n) \leqslant n+c$ for the constant $c=$ $f(1)-1$. Assume that $f(n) \neq n$ for some fixed $n$. When $m$ is large enough (e.g. $m \geqslant \max (n, c+1)$ ) we have

$$
m f(m)+n \leqslant m(m+c)+n \leqslant 2 m^{2}<2\left(m^{2}+f(n)\right)
$$

so we must have $m f(m)+n=m^{2}+f(n)$. This implies that

$$
0 \neq f(n)-n=m(f(m)-m)
$$

which is impossible for $m>|f(n)-n|$. It follows that $f$ is the identity function.

N2. Prove that for any pair of positive integers $k$ and $n$ there exist $k$ positive integers $m_{1}, m_{2}, \ldots, m_{k}$ such that

$$
1+\frac{2^{k}-1}{n}=\left(1+\frac{1}{m_{1}}\right)\left(1+\frac{1}{m_{2}}\right) \cdots\left(1+\frac{1}{m_{k}}\right) .
$$

(Japan)
Solution 1. We proceed by induction on $k$. For $k=1$ the statement is trivial. Assuming we have proved it for $k=j-1$, we now prove it for $k=j$.

Case 1. $n=2 t-1$ for some positive integer $t$.
Observe that

$$
1+\frac{2^{j}-1}{2 t-1}=\frac{2\left(t+2^{j-1}-1\right)}{2 t} \cdot \frac{2 t}{2 t-1}=\left(1+\frac{2^{j-1}-1}{t}\right)\left(1+\frac{1}{2 t-1}\right) .
$$

By the induction hypothesis we can find $m_{1}, \ldots, m_{j-1}$ such that

$$
1+\frac{2^{j-1}-1}{t}=\left(1+\frac{1}{m_{1}}\right)\left(1+\frac{1}{m_{2}}\right) \cdots\left(1+\frac{1}{m_{j-1}}\right)
$$

so setting $m_{j}=2 t-1$ gives the desired expression.
Case 2. $n=2 t$ for some positive integer $t$.
Now we have

$$
1+\frac{2^{j}-1}{2 t}=\frac{2 t+2^{j}-1}{2 t+2^{j}-2} \cdot \frac{2 t+2^{j}-2}{2 t}=\left(1+\frac{1}{2 t+2^{j}-2}\right)\left(1+\frac{2^{j-1}-1}{t}\right)
$$

noting that $2 t+2^{j}-2>0$. Again, we use that

$$
1+\frac{2^{j-1}-1}{t}=\left(1+\frac{1}{m_{1}}\right)\left(1+\frac{1}{m_{2}}\right) \cdots\left(1+\frac{1}{m_{j-1}}\right)
$$

Setting $m_{j}=2 t+2^{j}-2$ then gives the desired expression.
Solution 2. Consider the base 2 expansions of the residues of $n-1$ and $-n$ modulo $2^{k}$ :

$$
\begin{aligned}
n-1 & \equiv 2^{a_{1}}+2^{a_{2}}+\cdots+2^{a_{r}}\left(\bmod 2^{k}\right) & & \text { where } 0 \leqslant a_{1}<a_{2}<\ldots<a_{r} \leqslant k-1 \\
-n & \equiv 2^{b_{1}}+2^{b_{2}}+\cdots+2^{b_{s}}\left(\bmod 2^{k}\right) & & \text { where } 0 \leqslant b_{1}<b_{2}<\ldots<b_{s} \leqslant k-1 .
\end{aligned}
$$

Since $-1 \equiv 2^{0}+2^{1}+\cdots+2^{k-1}\left(\bmod 2^{k}\right)$, we have $\left\{a_{1}, \ldots, a_{r}\right\} \cup\left\{b_{1} \ldots, b_{s}\right\}=\{0,1, \ldots, k-1\}$ and $r+s=k$. Write

$$
\begin{aligned}
& S_{p}=2^{a_{p}}+2^{a_{p+1}}+\cdots+2^{a_{r}} \quad \text { for } 1 \leqslant p \leqslant r \\
& T_{q}=2^{b_{1}}+2^{b_{2}}+\cdots+2^{b_{q}} \quad \text { for } \quad 1 \leqslant q \leqslant s
\end{aligned}
$$

Also set $S_{r+1}=T_{0}=0$. Notice that $S_{1}+T_{s}=2^{k}-1$ and $n+T_{s} \equiv 0\left(\bmod 2^{k}\right)$. We have

$$
\begin{aligned}
1+\frac{2^{k}-1}{n} & =\frac{n+S_{1}+T_{s}}{n}=\frac{n+S_{1}+T_{s}}{n+T_{s}} \cdot \frac{n+T_{s}}{n} \\
& =\prod_{p=1}^{r} \frac{n+S_{p}+T_{s}}{n+S_{p+1}+T_{s}} \cdot \prod_{q=1}^{s} \frac{n+T_{q}}{n+T_{q-1}} \\
& =\prod_{p=1}^{r}\left(1+\frac{2^{a_{p}}}{n+S_{p+1}+T_{s}}\right) \cdot \prod_{q=1}^{s}\left(1+\frac{2^{b_{q}}}{n+T_{q-1}}\right)
\end{aligned}
$$

so if we define

$$
m_{p}=\frac{n+S_{p+1}+T_{s}}{2^{a_{p}}} \quad \text { for } 1 \leqslant p \leqslant r \quad \text { and } \quad m_{r+q}=\frac{n+T_{q-1}}{2^{b_{q}}} \quad \text { for } 1 \leqslant q \leqslant s
$$

the desired equality holds. It remains to check that every $m_{i}$ is an integer. For $1 \leqslant p \leqslant r$ we have

$$
n+S_{p+1}+T_{s} \equiv n+T_{s} \equiv 0 \quad\left(\bmod 2^{a_{p}}\right)
$$

and for $1 \leqslant q \leqslant r$ we have

$$
n+T_{q-1} \equiv n+T_{s} \equiv 0 \quad\left(\bmod 2^{b_{q}}\right)
$$

The desired result follows.

N3. Prove that there exist infinitely many positive integers $n$ such that the largest prime divisor of $n^{4}+n^{2}+1$ is equal to the largest prime divisor of $(n+1)^{4}+(n+1)^{2}+1$.
(Belgium)
Solution. Let $p_{n}$ be the largest prime divisor of $n^{4}+n^{2}+1$ and let $q_{n}$ be the largest prime divisor of $n^{2}+n+1$. Then $p_{n}=q_{n^{2}}$, and from

$$
n^{4}+n^{2}+1=\left(n^{2}+1\right)^{2}-n^{2}=\left(n^{2}-n+1\right)\left(n^{2}+n+1\right)=\left((n-1)^{2}+(n-1)+1\right)\left(n^{2}+n+1\right)
$$

it follows that $p_{n}=\max \left\{q_{n}, q_{n-1}\right\}$ for $n \geqslant 2$. Keeping in mind that $n^{2}-n+1$ is odd, we have

$$
\operatorname{gcd}\left(n^{2}+n+1, n^{2}-n+1\right)=\operatorname{gcd}\left(2 n, n^{2}-n+1\right)=\operatorname{gcd}\left(n, n^{2}-n+1\right)=1
$$

Therefore $q_{n} \neq q_{n-1}$.
To prove the result, it suffices to show that the set

$$
S=\left\{n \in \mathbb{Z}_{\geqslant 2} \mid q_{n}>q_{n-1} \text { and } q_{n}>q_{n+1}\right\}
$$

is infinite, since for each $n \in S$ one has

$$
p_{n}=\max \left\{q_{n}, q_{n-1}\right\}=q_{n}=\max \left\{q_{n}, q_{n+1}\right\}=p_{n+1} .
$$

Suppose on the contrary that $S$ is finite. Since $q_{2}=7<13=q_{3}$ and $q_{3}=13>7=q_{4}$, the set $S$ is non-empty. Since it is finite, we can consider its largest element, say $m$.

Note that it is impossible that $q_{m}>q_{m+1}>q_{m+2}>\ldots$ because all these numbers are positive integers, so there exists a $k \geqslant m$ such that $q_{k}<q_{k+1}$ (recall that $q_{k} \neq q_{k+1}$ ). Next observe that it is impossible to have $q_{k}<q_{k+1}<q_{k+2}<\ldots$, because $q_{(k+1)^{2}}=p_{k+1}=\max \left\{q_{k}, q_{k+1}\right\}=q_{k+1}$, so let us take the smallest $\ell \geqslant k+1$ such that $q_{\ell}>q_{\ell+1}$. By the minimality of $\ell$ we have $q_{\ell-1}<q_{\ell}$, so $\ell \in S$. Since $\ell \geqslant k+1>k \geqslant m$, this contradicts the maximality of $m$, and hence $S$ is indeed infinite.

Comment. Once the factorization of $n^{4}+n^{2}+1$ is found and the set $S$ is introduced, the problem is mainly about ruling out the case that

$$
\begin{equation*}
q_{k}<q_{k+1}<q_{k+2}<\ldots \tag{1}
\end{equation*}
$$

might hold for some $k \in \mathbb{Z}_{>0}$. In the above solution, this is done by observing $q_{(k+1)^{2}}=\max \left(q_{k}, q_{k+1}\right)$. Alternatively one may notice that (1) implies that $q_{j+2}-q_{j} \geqslant 6$ for $j \geqslant k+1$, since every prime greater than 3 is congruent to -1 or 1 modulo 6 . Then there is some integer $C \geqslant 0$ such that $q_{n} \geqslant 3 n-C$ for all $n \geqslant k$.

Now let the integer $t$ be sufficiently large (e.g. $t=\max \{k+1, C+3\}$ ) and set $p=q_{t-1} \geqslant 2 t$. Then $p \mid(t-1)^{2}+(t-1)+1$ implies that $p \mid(p-t)^{2}+(p-t)+1$, so $p$ and $q_{p-t}$ are prime divisors of $(p-t)^{2}+(p-t)+1$. But $p-t>t-1 \geqslant k$, so $q_{p-t}>q_{t-1}=p$ and $p \cdot q_{p-t}>p^{2}>(p-t)^{2}+(p-t)+1$, a contradiction.

N4. Determine whether there exists an infinite sequence of nonzero digits $a_{1}, a_{2}, a_{3}, \ldots$ and a positive integer $N$ such that for every integer $k>N$, the number $\overline{a_{k} a_{k-1} \ldots a_{1}}$ is a perfect square.
(Iran)
Answer. No.
Solution. Assume that $a_{1}, a_{2}, a_{3}, \ldots$ is such a sequence. For each positive integer $k$, let $y_{k}=$ $\overline{a_{k} a_{k-1} \ldots a_{1}}$. By the assumption, for each $k>N$ there exists a positive integer $x_{k}$ such that $y_{k}=x_{k}^{2}$.
I. For every $n$, let $5^{\gamma_{n}}$ be the greatest power of 5 dividing $x_{n}$. Let us show first that $2 \gamma_{n} \geqslant n$ for every positive integer $n>N$.

Assume, to the contrary, that there exists a positive integer $n>N$ such that $2 \gamma_{n}<n$, which yields

$$
y_{n+1}=\overline{a_{n+1} a_{n} \ldots a_{1}}=10^{n} a_{n+1}+\overline{a_{n} a_{n-1} \ldots a_{1}}=10^{n} a_{n+1}+y_{n}=5^{2 \gamma_{n}}\left(2^{n} 5^{n-2 \gamma_{n}} a_{n+1}+\frac{y_{n}}{5^{2 \gamma_{n}}}\right) .
$$

Since $5 \backslash y_{n} / 5^{2 \gamma_{n}}$, we obtain $\gamma_{n+1}=\gamma_{n}<n<n+1$. By the same arguments we obtain that $\gamma_{n}=\gamma_{n+1}=\gamma_{n+2}=\ldots$. Denote this common value by $\gamma$.

Now, for each $k \geqslant n$ we have

$$
\left(x_{k+1}-x_{k}\right)\left(x_{k+1}+x_{k}\right)=x_{k+1}^{2}-x_{k}^{2}=y_{k+1}-y_{k}=a_{k+1} \cdot 10^{k} .
$$

One of the numbers $x_{k+1}-x_{k}$ and $x_{k+1}+x_{k}$ is not divisible by $5^{\gamma+1}$ since otherwise one would have $5^{\gamma+1} \mid\left(\left(x_{k+1}-x_{k}\right)+\left(x_{k+1}+x_{k}\right)\right)=2 x_{k+1}$. On the other hand, we have $5^{k} \mid\left(x_{k+1}-x_{k}\right)\left(x_{k+1}+x_{k}\right)$, so $5^{k-\gamma}$ divides one of these two factors. Thus we get

$$
5^{k-\gamma} \leqslant \max \left\{x_{k+1}-x_{k}, x_{k+1}+x_{k}\right\}<2 x_{k+1}=2 \sqrt{y_{k+1}}<2 \cdot 10^{(k+1) / 2}
$$

which implies $5^{2 k}<4 \cdot 5^{2 \gamma} \cdot 10^{k+1}$, or $(5 / 2)^{k}<40 \cdot 5^{2 \gamma}$. The last inequality is clearly false for sufficiently large values of $k$. This contradiction shows that $2 \gamma_{n} \geqslant n$ for all $n>N$.
II. Consider now any integer $k>\max \{N / 2,2\}$. Since $2 \gamma_{2 k+1} \geqslant 2 k+1$ and $2 \gamma_{2 k+2} \geqslant 2 k+2$, we have $\gamma_{2 k+1} \geqslant k+1$ and $\gamma_{2 k+2} \geqslant k+1$. So, from $y_{2 k+2}=a_{2 k+2} \cdot 10^{2 k+1}+y_{2 k+1}$ we obtain $5^{2 k+2} \mid y_{2 k+2}-y_{2 k+1}=a_{2 k+2} \cdot 10^{2 k+1}$ and thus $5 \mid a_{2 k+2}$, which implies $a_{2 k+2}=5$. Therefore,

$$
\left(x_{2 k+2}-x_{2 k+1}\right)\left(x_{2 k+2}+x_{2 k+1}\right)=x_{2 k+2}^{2}-x_{2 k+1}^{2}=y_{2 k+2}-y_{2 k+1}=5 \cdot 10^{2 k+1}=2^{2 k+1} \cdot 5^{2 k+2} .
$$

Setting $A_{k}=x_{2 k+2} / 5^{k+1}$ and $B_{k}=x_{2 k+1} / 5^{k+1}$, which are integers, we obtain

$$
\begin{equation*}
\left(A_{k}-B_{k}\right)\left(A_{k}+B_{k}\right)=2^{2 k+1} \tag{1}
\end{equation*}
$$

Both $A_{k}$ and $B_{k}$ are odd, since otherwise $y_{2 k+2}$ or $y_{2 k+1}$ would be a multiple of 10 which is false by $a_{1} \neq 0$; so one of the numbers $A_{k}-B_{k}$ and $A_{k}+B_{k}$ is not divisible by 4. Therefore (1) yields $A_{k}-B_{k}=2$ and $A_{k}+B_{k}=2^{2 k}$, hence $A_{k}=2^{2 k-1}+1$ and thus

$$
x_{2 k+2}=5^{k+1} A_{k}=10^{k+1} \cdot 2^{k-2}+5^{k+1}>10^{k+1}
$$

since $k \geqslant 2$. This implies that $y_{2 k+2}>10^{2 k+2}$ which contradicts the fact that $y_{2 k+2}$ contains $2 k+2$ digits. The desired result follows.

Solution 2. Again, we assume that a sequence $a_{1}, a_{2}, a_{3}, \ldots$ satisfies the problem conditions, introduce the numbers $x_{k}$ and $y_{k}$ as in the previous solution, and notice that

$$
\begin{equation*}
y_{k+1}-y_{k}=\left(x_{k+1}-x_{k}\right)\left(x_{k+1}+x_{k}\right)=10^{k} a_{k+1} \tag{2}
\end{equation*}
$$

for all $k>N$. Consider any such $k$. Since $a_{1} \neq 0$, the numbers $x_{k}$ and $x_{k+1}$ are not multiples of 10, and therefore the numbers $p_{k}=x_{k+1}-x_{k}$ and $q_{k}=x_{k+1}+x_{k}$ cannot be simultaneously multiples of 20 , and hence one of them is not divisible either by 4 or by 5 . In view of (2), this means that the other one is divisible by either $5^{k}$ or by $2^{k-1}$. Notice also that $p_{k}$ and $q_{k}$ have the same parity, so both are even.

On the other hand, we have $x_{k+1}^{2}=x_{k}^{2}+10^{k} a_{k+1} \geqslant x_{k}^{2}+10^{k}>2 x_{k}^{2}$, so $x_{k+1} / x_{k}>\sqrt{2}$, which implies that

$$
\begin{equation*}
1<\frac{q_{k}}{p_{k}}=1+\frac{2}{x_{k+1} / x_{k}-1}<1+\frac{2}{\sqrt{2}-1}<6 . \tag{3}
\end{equation*}
$$

Thus, if one of the numbers $p_{k}$ and $q_{k}$ is divisible by $5^{k}$, then we have

$$
10^{k+1}>10^{k} a_{k+1}=p_{k} q_{k} \geqslant \frac{\left(5^{k}\right)^{2}}{6}
$$

and hence $(5 / 2)^{k}<60$ which is false for sufficiently large $k$. So, assuming that $k$ is large, we get that $2^{k-1}$ divides one of the numbers $p_{k}$ and $q_{k}$. Hence
$\left\{p_{k}, q_{k}\right\}=\left\{2^{k-1} \cdot 5^{r_{k}} b_{k}, 2 \cdot 5^{k-r_{k}} c_{k}\right\} \quad$ with nonnegative integers $b_{k}, c_{k}, r_{k}$ such that $b_{k} c_{k}=a_{k+1}$.
Moreover, from (3) we get

$$
6>\frac{2^{k-1} \cdot 5^{r_{k}} b_{k}}{2 \cdot 5^{k-r_{k}} c_{k}} \geqslant \frac{1}{36} \cdot\left(\frac{2}{5}\right)^{k} \cdot 5^{2 r_{k}} \quad \text { and } \quad 6>\frac{2 \cdot 5^{k-r_{k}} c_{k}}{2^{k-1} \cdot 5^{r_{k}} b_{k}} \geqslant \frac{4}{9} \cdot\left(\frac{5}{2}\right)^{k} \cdot 5^{-2 r_{k}}
$$

so

$$
\begin{equation*}
\alpha k+c_{1}<r_{k}<\alpha k+c_{2} \quad \text { for } \alpha=\frac{1}{2} \log _{5}\left(\frac{5}{2}\right)<1 \text { and some constants } c_{2}>c_{1} . \tag{4}
\end{equation*}
$$

Consequently, for $C=c_{2}-c_{1}+1-\alpha>0$ we have

$$
\begin{equation*}
(k+1)-r_{k+1} \leqslant k-r_{k}+C \tag{5}
\end{equation*}
$$

Next, we will use the following easy lemma.
Lemma. Let $s$ be a positive integer. Then $5^{s+2^{s}} \equiv 5^{s}\left(\bmod 10^{s}\right)$.
Proof. Euler's theorem gives $5^{2^{s}} \equiv 1\left(\bmod 2^{s}\right)$, so $5^{s+2^{s}}-5^{s}=5^{s}\left(5^{2^{s}}-1\right)$ is divisible by $2^{s}$ and $5^{s}$.
Now, for every large $k$ we have

$$
\begin{equation*}
x_{k+1}=\frac{p_{k}+q_{k}}{2}=5^{r_{k}} \cdot 2^{k-2} b_{k}+5^{k-r_{k}} c_{k} \equiv 5^{k-r_{k}} c_{k} \quad\left(\bmod 10^{r_{k}}\right) \tag{6}
\end{equation*}
$$

since $r_{k} \leqslant k-2$ by (4); hence $y_{k+1} \equiv 5^{2\left(k-r_{k}\right)} c_{k}^{2}\left(\bmod 10^{r_{k}}\right)$. Let us consider some large integer $s$, and choose the minimal $k$ such that $2\left(k-r_{k}\right) \geqslant s+2^{s}$; it exists by (4). Set $d=2\left(k-r_{k}\right)-\left(s+2^{s}\right)$. By (4) we have $2^{s}<2\left(k-r_{k}\right)<\left(\frac{2}{\alpha}-2\right) r_{k}-\frac{2 c_{1}}{\alpha}$; if $s$ is large this implies $r_{k}>s$, so (6) also holds modulo $10^{s}$. Then (6) and the lemma give

$$
\begin{equation*}
y_{k+1} \equiv 5^{2\left(k-r_{k}\right)} c_{k}^{2}=5^{s+2^{s}} \cdot 5^{d} c_{k}^{2} \equiv 5^{s} \cdot 5^{d} c_{k}^{2} \quad\left(\bmod 10^{s}\right) . \tag{7}
\end{equation*}
$$

By (5) and the minimality of $k$ we have $d \leqslant 2 C$, so $5^{d} c_{k}^{2} \leqslant 5^{2 C} \cdot 81=D$. Using $5^{4}<10^{3}$ we obtain

$$
5^{s} \cdot 5^{d} c_{k}^{2}<10^{3 s / 4} D<10^{s-1}
$$

for sufficiently large $s$. This, together with (7), shows that the $s$ th digit from the right in $y_{k+1}$, which is $a_{s}$, is zero. This contradicts the problem condition.

N5. Fix an integer $k \geqslant 2$. Two players, called Ana and Banana, play the following game of numbers: Initially, some integer $n \geqslant k$ gets written on the blackboard. Then they take moves in turn, with Ana beginning. A player making a move erases the number $m$ just written on the blackboard and replaces it by some number $m^{\prime}$ with $k \leqslant m^{\prime}<m$ that is coprime to $m$. The first player who cannot move anymore loses.

An integer $n \geqslant k$ is called good if Banana has a winning strategy when the initial number is $n$, and bad otherwise.

Consider two integers $n, n^{\prime} \geqslant k$ with the property that each prime number $p \leqslant k$ divides $n$ if and only if it divides $n^{\prime}$. Prove that either both $n$ and $n^{\prime}$ are good or both are bad.
(Italy)
Solution 1. Let us first observe that the number appearing on the blackboard decreases after every move; so the game necessarily ends after at most $n$ steps, and consequently there always has to be some player possessing a winning strategy. So if some $n \geqslant k$ is bad, then Ana has a winning strategy in the game with starting number $n$.

More precisely, if $n \geqslant k$ is such that there is a good integer $m$ with $n>m \geqslant k$ and $\operatorname{gcd}(m, n)=1$, then $n$ itself is bad, for Ana has the following winning strategy in the game with initial number $n$ : She proceeds by first playing $m$ and then using Banana's strategy for the game with starting number $m$.

Otherwise, if some integer $n \geqslant k$ has the property that every integer $m$ with $n>m \geqslant k$ and $\operatorname{gcd}(m, n)=1$ is bad, then $n$ is good. Indeed, if Ana can make a first move at all in the game with initial number $n$, then she leaves it in a position where the first player has a winning strategy, so that Banana can defeat her.

In particular, this implies that any two good numbers have a non-trivial common divisor. Also, $k$ itself is good.

For brevity, we say that $n \longrightarrow x$ is a move if $n$ and $x$ are two coprime integers with $n>x \geqslant k$.
Claim 1. If $n$ is good and $n^{\prime}$ is a multiple of $n$, then $n^{\prime}$ is also good.
Proof. If $n^{\prime}$ were bad, there would have to be some move $n^{\prime} \longrightarrow x$, where $x$ is good. As $n^{\prime}$ is a multiple of $n$ this implies that the two good numbers $n$ and $x$ are coprime, which is absurd.

Claim 2. If $r$ and $s$ denote two positive integers for which $r s \geqslant k$ is bad, then $r^{2} s$ is also bad. Proof. Since $r s$ is bad, there is a move $r s \longrightarrow x$ for some good $x$. Evidently $x$ is coprime to $r^{2} s$ as well, and hence the move $r^{2} s \longrightarrow x$ shows that $r^{2} s$ is indeed bad.

Claim 3. If $p>k$ is prime and $n \geqslant k$ is bad, then np is also bad.
Proof. Otherwise we choose a counterexample with $n$ being as small as possible. In particular, $n p$ is good. Since $n$ is bad, there is a move $n \longrightarrow x$ for some good $x$. Now $n p \longrightarrow x$ cannot be a valid move, which tells us that $x$ has to be divisible by $p$. So we can write $x=p^{r} y$, where $r$ and $y$ denote some positive integers, the latter of which is not divisible by $p$.

Note that $y=1$ is impossible, for then we would have $x=p^{r}$ and the move $x \longrightarrow k$ would establish that $x$ is bad. In view of this, there is a least power $y^{\alpha}$ of $y$ that is at least as large as $k$. Since the numbers $n p$ and $y^{\alpha}$ are coprime and the former is good, the latter has to be bad. Moreover, the minimality of $\alpha$ implies $y^{\alpha}<k y<p y=\frac{x}{p^{r-1}}<\frac{n}{p^{r-1}}$. So $p^{r-1} \cdot y^{\alpha}<n$ and consequently all the numbers $y^{\alpha}, p y^{\alpha}, \ldots, p^{r} \cdot y^{\alpha}=p\left(p^{r-1} \cdot y^{\alpha}\right)$ are bad due to the minimal choice of $n$. But now by Claim 1 the divisor $x$ of $p^{r} \cdot y^{\alpha}$ cannot be good, whereby we have reached a contradiction that proves Claim 3.

We now deduce the statement of the problem from these three claims. To this end, we call two integers $a, b \geqslant k$ similar if they are divisible by the same prime numbers not exceeding $k$. We are to prove that if $a$ and $b$ are similar, then either both of them are good or both are bad. As in this case the product $a b$ is similar to both $a$ and $b$, it suffices to show the following: if $c \geqslant k$ is similar to some of its multiples $d$, then either both $c$ and $d$ are good or both are bad.

Assuming that this is not true in general, we choose a counterexample $\left(c_{0}, d_{0}\right)$ with $d_{0}$ being as small as possible. By Claim 1, $c_{0}$ is bad whilst $d_{0}$ is good. Plainly $d_{0}$ is strictly greater than $c_{0}$ and hence the quotient $\frac{d_{0}}{c_{0}}$ has some prime factor $p$. Clearly $p$ divides $d_{0}$. If $p \leqslant k$, then $p$ divides $c_{0}$ as well due to the similarity, and hence $d_{0}$ is actually divisible by $p^{2}$. So $\frac{d_{0}}{p}$ is good by the contrapositive of Claim 2. Since $c_{0} \left\lvert\, \frac{d_{0}}{p}\right.$, the pair ( $c_{0}, \frac{d_{0}}{p}$ ) contradicts the supposed minimality of $d_{0}$. This proves $p>k$, but now we get the same contradiction using Claim 3 instead of Claim 2 . Thereby the problem is solved.

Solution 2. We use the same analysis of the game of numbers as in the first five paragraphs of the first solution. Let us call a prime number $p$ small in case $p \leqslant k$ and big otherwise. We again call two integers similar if their sets of small prime factors coincide.

Claim 4. For each integer $b \geqslant k$ having some small prime factor, there exists an integer $x$ similar to it with $b \geqslant x \geqslant k$ and having no big prime factors.
Proof. Unless $b$ has a big prime factor we may simply choose $x=b$. Now let $p$ and $q$ denote a small and a big prime factor of $b$, respectively. Let $a$ be the product of all small prime factors of $b$. Further define $n$ to be the least non-negative integer for which the number $x=p^{n} a$ is at least as large as $k$. It suffices to show that $b>x$. This is clear in case $n=0$, so let us assume $n>0$ from now on. Then we have $x<p k$ due to the minimality of $n, p \leqslant a$ because $p$ divides $a$ by construction, and $k<q$. Therefore $x<a q$ and, as the right hand side is a product of distinct prime factors of $b$, this implies indeed $x<b$.

Let us now assume that there is a pair $(a, b)$ of similar numbers such that $a$ is bad and $b$ is good. Take such a pair with $\max (a, b)$ being as small as possible. Since $a$ is bad, there exists a move $a \longrightarrow r$ for some good $r$. Since the numbers $k$ and $r$ are both good, they have a common prime factor, which necessarily has to be small. Thus Claim 4 is applicable to $r$, which yields an integer $r^{\prime}$ similar to $r$ containing small prime factors only and satisfying $r \geqslant r^{\prime} \geqslant k$. Since $\max \left(r, r^{\prime}\right)=r<a \leqslant \max (a, b)$ the number $r^{\prime}$ is also good. Now let $p$ denote a common prime factor of the good numbers $r^{\prime}$ and $b$. By our construction of $r^{\prime}$, this prime is small and due to the similarities it consequently divides $a$ and $r$, contrary to $a \longrightarrow r$ being a move. Thereby the problem is solved.

Comment 1. Having reached Claim 4 of Solution 2, there are various other ways to proceed. For instance, one may directly obtain the following fact, which seems to be interesting in its own right:

Claim 5. Any two good numbers have a common small prime factor.
Proof. Otherwise there exists a pair $\left(b, b^{\prime}\right)$ of good numbers with $b^{\prime} \geqslant b \geqslant k$ all of whose common prime factors are big. Choose such a pair with $b^{\prime}$ being as small as possible. Since $b$ and $k$ are both good, there has to be a common prime factor $p$ of $b$ and $k$. Evidently $p$ is small and thus it cannot divide $b^{\prime}$, which in turn tells us $b^{\prime}>b$. Applying Claim 4 to $b$ we get an integer $x$ with $b \geqslant x \geqslant k$ that is similar to $b$ and has no big prime divisors at all. By our assumption, $b^{\prime}$ and $x$ are coprime, and as $b^{\prime}$ is good this implies that $x$ is bad. Consequently there has to be some move $x \longrightarrow b^{*}$ such that $b^{*}$ is good. But now all the small prime factors of $b$ also appear in $x$ and thus they cannot divide $b^{*}$. Therefore the pair $\left(b^{*}, b\right)$ contradicts the supposed minimality of $b^{\prime}$.

From that point, it is easy to complete the solution: assume that there are two similar integers $a$ and $b$ such that $a$ is bad and $b$ is good. Since $a$ is bad, there is a move $a \longrightarrow b^{\prime}$ for some good $b^{\prime}$. By Claim 5 , there is a small prime $p$ dividing $b$ and $b^{\prime}$. Due to the similarity of $a$ and $b$, the prime $p$ has to divide $a$ as well, but this contradicts the fact that $a \longrightarrow b^{\prime}$ is a valid move. Thereby the problem is solved.

Comment 2. There are infinitely many good numbers, e.g. all multiples of $k$. The increasing sequence $b_{0}, b_{1}, \ldots$, of all good numbers may be constructed recursively as follows:

- Start with $b_{0}=k$.
- If $b_{n}$ has just been defined for some $n \geqslant 0$, then $b_{n+1}$ is the smallest number $b>b_{n}$ that is coprime to none of $b_{0}, \ldots, b_{n}$.

This construction can be used to determine the set of good numbers for any specific $k$ as explained in the next comment. It is already clear that if $k=p^{\alpha}$ is a prime power, then a number $b \geqslant k$ is good if and only if it is divisible by $p$.

Comment 3. Let $P>1$ denote the product of all small prime numbers. Then any two integers $a, b \geqslant k$ that are congruent modulo $P$ are similar. Thus the infinite word $W_{k}=\left(X_{k}, X_{k+1}, \ldots\right)$ defined by

$$
X_{i}= \begin{cases}A & \text { if } i \text { is bad } \\ B & \text { if } i \text { is good }\end{cases}
$$

for all $i \geqslant k$ is periodic and the length of its period divides $P$. As the prime power example shows, the true period can sometimes be much smaller than $P$. On the other hand, there are cases where the period is rather large; e.g., if $k=15$, the sequence of good numbers begins with $15,18,20,24,30,36,40,42,45$ and the period of $W_{15}$ is 30 .

Comment 4. The original proposal contained two questions about the game of numbers, namely (a) to show that if two numbers have the same prime factors then either both are good or both are bad, and (b) to show that the word $W_{k}$ introduced in the previous comment is indeed periodic. The Problem Selection Committee thinks that the above version of the problem is somewhat easier, even though it demands to prove a stronger result.

N6. Determine all functions $f: \mathbb{Q} \longrightarrow \mathbb{Z}$ satisfying

$$
\begin{equation*}
f\left(\frac{f(x)+a}{b}\right)=f\left(\frac{x+a}{b}\right) \tag{1}
\end{equation*}
$$

for all $x \in \mathbb{Q}, a \in \mathbb{Z}$, and $b \in \mathbb{Z}_{>0}$. (Here, $\mathbb{Z}_{>0}$ denotes the set of positive integers.)

Answer. There are three kinds of such functions, which are: all constant functions, the floor function, and the ceiling function.
Solution 1. I. We start by verifying that these functions do indeed satisfy (1). This is clear for all constant functions. Now consider any triple $(x, a, b) \in \mathbb{Q} \times \mathbb{Z} \times \mathbb{Z}_{>0}$ and set

$$
q=\left\lfloor\frac{x+a}{b}\right\rfloor .
$$

This means that $q$ is an integer and $b q \leqslant x+a<b(q+1)$. It follows that $b q \leqslant\lfloor x\rfloor+a<b(q+1)$ holds as well, and thus we have

$$
\left\lfloor\frac{\lfloor x\rfloor+a}{b}\right\rfloor=\left\lfloor\frac{x+a}{b}\right\rfloor,
$$

meaning that the floor function does indeed satisfy (1). One can check similarly that the ceiling function has the same property.
II. Let us now suppose conversely that the function $f: \mathbb{Q} \longrightarrow \mathbb{Z}$ satisfies (1) for all $(x, a, b) \in$ $\mathbb{Q} \times \mathbb{Z} \times \mathbb{Z}_{>0}$. According to the behaviour of the restriction of $f$ to the integers we distinguish two cases.

Case 1: There is some $m \in \mathbb{Z}$ such that $f(m) \neq m$.
Write $f(m)=C$ and let $\eta \in\{-1,+1\}$ and $b$ denote the sign and absolute value of $f(m)-m$, respectively. Given any integer $r$, we may plug the triple ( $m, r b-C, b$ ) into (1), thus getting $f(r)=f(r-\eta)$. Starting with $m$ and using induction in both directions, we deduce from this that the equation $f(r)=C$ holds for all integers $r$. Now any rational number $y$ can be written in the form $y=\frac{p}{q}$ with $(p, q) \in \mathbb{Z} \times \mathbb{Z}_{>0}$, and substituting $(C-p, p-C, q)$ into (1) we get $f(y)=f(0)=C$. Thus $f$ is the constant function whose value is always $C$.

Case 2: One has $f(m)=m$ for all integers $m$.
Note that now the special case $b=1$ of (1) takes a particularly simple form, namely

$$
\begin{equation*}
f(x)+a=f(x+a) \quad \text { for all }(x, a) \in \mathbb{Q} \times \mathbb{Z} . \tag{2}
\end{equation*}
$$

Defining $f\left(\frac{1}{2}\right)=\omega$ we proceed in three steps.
Step $A$. We show that $\omega \in\{0,1\}$.
If $\omega \leqslant 0$, we may plug $\left(\frac{1}{2},-\omega, 1-2 \omega\right)$ into (1), obtaining $0=f(0)=f\left(\frac{1}{2}\right)=\omega$. In the contrary case $\omega \geqslant 1$ we argue similarly using the triple $\left(\frac{1}{2}, \omega-1,2 \omega-1\right)$.

Step B. We show that $f(x)=\omega$ for all rational numbers $x$ with $0<x<1$.
Assume that this fails and pick some rational number $\frac{a}{b} \in(0,1)$ with minimal $b$ such that $f\left(\frac{a}{b}\right) \neq \omega$. Obviously, $\operatorname{gcd}(a, b)=1$ and $b \geqslant 2$. If $b$ is even, then $a$ has to be odd and we can substitute $\left(\frac{1}{2}, \frac{a-1}{2}, \frac{b}{2}\right)$ into (1), which yields

$$
\begin{equation*}
f\left(\frac{\omega+(a-1) / 2}{b / 2}\right)=f\left(\frac{a}{b}\right) \neq \omega . \tag{3}
\end{equation*}
$$

Recall that $0 \leqslant(a-1) / 2<b / 2$. Thus, in both cases $\omega=0$ and $\omega=1$, the left-hand part of (3) equals $\omega$ either by the minimality of $b$, or by $f(\omega)=\omega$. A contradiction.

Thus $b$ has to be odd, so $b=2 k+1$ for some $k \geqslant 1$. Applying (1) to $\left(\frac{1}{2}, k, b\right)$ we get

$$
\begin{equation*}
f\left(\frac{\omega+k}{b}\right)=f\left(\frac{1}{2}\right)=\omega \tag{4}
\end{equation*}
$$

Since $a$ and $b$ are coprime, there exist integers $r \in\{1,2, \ldots, b\}$ and $m$ such that $r a-m b=k+\omega$. Note that we actually have $1 \leqslant r<b$, since the right hand side is not a multiple of $b$. If $m$ is negative, then we have $r a-m b>b \geqslant k+\omega$, which is absurd. Similarly, $m \geqslant r$ leads to $r a-m b<b r-b r=0$, which is likewise impossible; so we must have $0 \leqslant m \leqslant r-1$.

We finally substitute $\left(\frac{k+\omega}{b}, m, r\right)$ into (1) and use (4) to learn

$$
f\left(\frac{\omega+m}{r}\right)=f\left(\frac{a}{b}\right) \neq \omega .
$$

But as above one may see that the left hand side has to equal $\omega$ due to the minimality of $b$. This contradiction concludes our step B.

Step $C$. Now notice that if $\omega=0$, then $f(x)=\lfloor x\rfloor$ holds for all rational $x$ with $0 \leqslant x<1$ and hence by (2) this even holds for all rational numbers $x$. Similarly, if $\omega=1$, then $f(x)=\lceil x\rceil$ holds for all $x \in \mathbb{Q}$. Thereby the problem is solved.

Comment 1. An alternative treatment of Steps B and C from the second case, due to the proposer, proceeds as follows. Let square brackets indicate the floor function in case $\omega=0$ and the ceiling function if $\omega=1$. We are to prove that $f(x)=[x]$ holds for all $x \in \mathbb{Q}$, and because of Step A and (2) we already know this in case $2 x \in \mathbb{Z}$. Applying (1) to $(2 x, 0,2)$ we get

$$
f(x)=f\left(\frac{f(2 x)}{2}\right)
$$

and by the previous observation this yields

$$
\begin{equation*}
f(x)=\left[\frac{f(2 x)}{2}\right] \quad \text { for all } x \in \mathbb{Q} \tag{5}
\end{equation*}
$$

An easy induction now shows

$$
\begin{equation*}
f(x)=\left[\frac{f\left(2^{n} x\right)}{2^{n}}\right] \quad \text { for all }(x, n) \in \mathbb{Q} \times \mathbb{Z}_{>0} \tag{6}
\end{equation*}
$$

Now suppose first that $x$ is not an integer but can be written in the form $\frac{p}{q}$ with $p \in \mathbb{Z}$ and $q \in \mathbb{Z}_{>0}$ both being odd. Let $d$ denote the multiplicative order of 2 modulo $q$ and let $m$ be any large integer. Plugging $n=d m$ into (6) and using (2) we get

$$
f(x)=\left[\frac{f\left(2^{d m} x\right)}{2^{d m}}\right]=\left[\frac{f(x)+\left(2^{d m}-1\right) x}{2^{d m}}\right]=\left[x+\frac{f(x)-x}{2^{d m}}\right] .
$$

Since $x$ is not an integer, the square bracket function is continuous at $x$; hence as $m$ tends to infinity the above fomula gives $f(x)=[x]$. To complete the argument we just need to observe that if some $y \in \mathbb{Q}$ satisfies $f(y)=[y]$, then (5) yields $f\left(\frac{y}{2}\right)=f\left(\frac{[y]}{2}\right)=\left[\frac{[y]}{2}\right]=\left[\frac{y}{2}\right]$.

Solution 2. Here we just give another argument for the second case of the above solution. Again we use equation (2). It follows that the set $S$ of all zeros of $f$ contains for each $x \in \mathbb{Q}$ exactly one term from the infinite sequence $\ldots, x-2, x-1, x, x+1, x+2, \ldots$.

Next we claim that

$$
\begin{equation*}
\text { if }(p, q) \in \mathbb{Z} \times \mathbb{Z}_{>0} \text { and } \frac{p}{q} \in S \text {, then } \frac{p}{q+1} \in S \text { holds as well. } \tag{7}
\end{equation*}
$$

To see this we just plug $\left(\frac{p}{q}, p, q+1\right)$ into (1), thus getting $f\left(\frac{p}{q+1}\right)=f\left(\frac{p}{q}\right)=0$.
From this we get that

$$
\begin{equation*}
\text { if } x, y \in \mathbb{Q}, x>y>0, \text { and } x \in S, \text { then } y \in S . \tag{8}
\end{equation*}
$$

Indeed, if we write $x=\frac{p}{q}$ and $y=\frac{r}{s}$ with $p, q, r, s \in \mathbb{Z}_{>0}$, then $p s>q r$ and (7) tells us

$$
0=f\left(\frac{p}{q}\right)=f\left(\frac{p r}{q r}\right)=f\left(\frac{p r}{q r+1}\right)=\ldots=f\left(\frac{p r}{p s}\right)=f\left(\frac{r}{s}\right) .
$$

Essentially the same argument also establishes that

$$
\begin{equation*}
\text { if } x, y \in \mathbb{Q}, x<y<0, \text { and } x \in S, \text { then } y \in S \tag{9}
\end{equation*}
$$

From (8) and (9) we get $0 \in S \subseteq(-1,+1)$ and hence the real number $\alpha=\sup (S)$ exists and satisfies $0 \leqslant \alpha \leqslant 1$.

Let us assume that we actually had $0<\alpha<1$. Note that $f(x)=0$ if $x \in(0, \alpha) \cap \mathbb{Q}$ by (8), and $f(x)=1$ if $x \in(\alpha, 1) \cap \mathbb{Q}$ by (9) and (2). Let $K$ denote the unique positive integer satisfying $K \alpha<1 \leqslant(K+1) \alpha$. The first of these two inequalities entails $\alpha<\frac{1+\alpha}{K+1}$, and thus there is a rational number $x \in\left(\alpha, \frac{1+\alpha}{K+1}\right)$. Setting $y=(K+1) x-1$ and substituting $(y, 1, K+1)$ into (1) we learn

$$
f\left(\frac{f(y)+1}{K+1}\right)=f\left(\frac{y+1}{K+1}\right)=f(x) .
$$

Since $\alpha<x<1$ and $0<y<\alpha$, this simplifies to

$$
f\left(\frac{1}{K+1}\right)=1
$$

But, as $0<\frac{1}{K+1} \leqslant \alpha$, this is only possible if $\alpha=\frac{1}{K+1}$ and $f(\alpha)=1$. From this, however, we get the contradiction

$$
0=f\left(\frac{1}{(K+1)^{2}}\right)=f\left(\frac{\alpha+0}{K+1}\right)=f\left(\frac{f(\alpha)+0}{K+1}\right)=f(\alpha)=1
$$

Thus our assumption $0<\alpha<1$ has turned out to be wrong and it follows that $\alpha \in\{0,1\}$. If $\alpha=0$, then we have $S \subseteq(-1,0]$, whence $S=(-1,0] \cap \mathbb{Q}$, which in turn yields $f(x)=\lceil x\rceil$ for all $x \in \mathbb{Q}$ due to (2). Similarly, $\alpha=1$ entails $S=[0,1) \cap \mathbb{Q}$ and $f(x)=\lfloor x\rfloor$ for all $x \in \mathbb{Q}$. Thereby the solution is complete.

Comment 2. It seems that all solutions to this problems involve some case distinction separating the constant solutions from the unbounded ones, though the "descriptions" of the cases may be different depending on the work that has been done at the beginning of the solution. For instance, these two cases can also be " $f$ is periodic on the integers" and " $f$ is not periodic on the integers". The case leading to the unbounded solutions appears to be the harder one.

In most approaches, the cases leading to the two functions $x \longmapsto\lfloor x\rfloor$ and $x \longmapsto\lceil x\rceil$ can easily be treated parallelly, but sometimes it may be useful to know that there is some symmetry in the problem interchanging these two functions. Namely, if a function $f: \mathbb{Q} \longrightarrow \mathbb{Z}$ satisfies (1), then so does the function $g: \mathbb{Q} \longrightarrow \mathbb{Z}$ defined by $g(x)=-f(-x)$ for all $x \in \mathbb{Q}$. For that reason, we could have restricted our attention to the case $\omega=0$ in the first solution and, once $\alpha \in\{0,1\}$ had been obtained, to the case $\alpha=0$ in the second solution.

N7. Let $\nu$ be an irrational positive number, and let $m$ be a positive integer. A pair $(a, b)$ of positive integers is called good if

$$
\begin{equation*}
a\lceil b \nu\rceil-b\lfloor a \nu\rfloor=m \tag{*}
\end{equation*}
$$

A good pair $(a, b)$ is called excellent if neither of the pairs $(a-b, b)$ and $(a, b-a)$ is good. (As usual, by $\lfloor x\rfloor$ and $\lceil x\rceil$ we denote the integer numbers such that $x-1<\lfloor x\rfloor \leqslant x$ and $x \leqslant\lceil x\rceil<x+1$.)

Prove that the number of excellent pairs is equal to the sum of the positive divisors of $m$.
(U.S.A.)

Solution. For positive integers $a$ and $b$, let us denote

$$
f(a, b)=a\lceil b \nu\rceil-b\lfloor a \nu\rfloor .
$$

We will deal with various values of $m$; thus it is convenient to say that a pair $(a, b)$ is $m$-good or $m$-excellent if the corresponding conditions are satisfied.

To start, let us investigate how the values $f(a+b, b)$ and $f(a, b+a)$ are related to $f(a, b)$. If $\{a \nu\}+\{b \nu\}<1$, then we have $\lfloor(a+b) \nu\rfloor=\lfloor a \nu\rfloor+\lfloor b \nu\rfloor$ and $\lceil(a+b) \nu\rceil=\lceil a \nu\rceil+\lceil b \nu\rceil-1$, so

$$
f(a+b, b)=(a+b)\lceil b \nu\rceil-b(\lfloor a \nu\rfloor+\lfloor b \nu\rfloor)=f(a, b)+b(\lceil b \nu\rceil-\lfloor b \nu\rfloor)=f(a, b)+b
$$

and

$$
f(a, b+a)=a(\lceil b \nu\rceil+\lceil a \nu\rceil-1)-(b+a)\lfloor a \nu\rfloor=f(a, b)+a(\lceil a \nu\rceil-1-\lfloor a \nu\rfloor)=f(a, b) .
$$

Similarly, if $\{a \nu\}+\{b \nu\} \geqslant 1$ then one obtains

$$
f(a+b, b)=f(a, b) \quad \text { and } \quad f(a, b+a)=f(a, b)+a .
$$

So, in both cases one of the numbers $f(a+b, a)$ and $f(a, b+a)$ is equal to $f(a, b)$ while the other is greater than $f(a, b)$ by one of $a$ and $b$. Thus, exactly one of the pairs $(a+b, b)$ and $(a, b+a)$ is excellent (for an appropriate value of $m$ ).

Now let us say that the pairs $(a+b, b)$ and $(a, b+a)$ are the children of the pair $(a, b)$, while this pair is their parent. Next, if a pair $(c, d)$ can be obtained from $(a, b)$ by several passings from a parent to a child, we will say that $(c, d)$ is a descendant of $(a, b)$, while $(a, b)$ is an ancestor of $(c, d)$ (a pair is neither an ancestor nor a descendant of itself). Thus each pair ( $a, b$ ) has two children, it has a unique parent if $a \neq b$, and no parents otherwise. Therefore, each pair of distinct positive integers has a unique ancestor of the form $(a, a)$; our aim is now to find how many $m$-excellent descendants each such pair has.

Notice now that if a pair $(a, b)$ is $m$-excellent then $\min \{a, b\} \leqslant m$. Indeed, if $a=b$ then $f(a, a)=a=m$, so the statement is valid. Otherwise, the pair $(a, b)$ is a child of some pair $\left(a^{\prime}, b^{\prime}\right)$. If $b=b^{\prime}$ and $a=a^{\prime}+b^{\prime}$, then we should have $m=f(a, b)=f\left(a^{\prime}, b^{\prime}\right)+b^{\prime}$, so $b=b^{\prime}=m-f\left(a^{\prime}, b^{\prime}\right)<m$. Similarly, if $a=a^{\prime}$ and $b=b^{\prime}+a^{\prime}$ then $a<m$.

Let us consider the set $S_{m}$ of all pairs $(a, b)$ such that $f(a, b) \leqslant m$ and $\min \{a, b\} \leqslant m$. Then all the ancestors of the elements in $S_{m}$ are again in $S_{m}$, and each element in $S_{m}$ either is of the form ( $a, a$ ) with $a \leqslant m$, or has a unique ancestor of this form. From the arguments above we see that all $m$-excellent pairs lie in $S_{m}$.

We claim now that the set $S_{m}$ is finite. Indeed, assume, for instance, that it contains infinitely many pairs ( $c, d$ ) with $d>2 m$. Such a pair is necessarily a child of $(c, d-c)$, and thus a descendant of some pair $\left(c, d^{\prime}\right)$ with $m<d^{\prime} \leqslant 2 m$. Therefore, one of the pairs $(a, b) \in S_{m}$ with $m<b \leqslant 2 m$
has infinitely many descendants in $S_{m}$, and all these descendants have the form $(a, b+k a)$ with $k$ a positive integer. Since $f(a, b+k a)$ does not decrease as $k$ grows, it becomes constant for $k \geqslant k_{0}$, where $k_{0}$ is some positive integer. This means that $\{a \nu\}+\{(b+k a) \nu\}<1$ for all $k \geqslant k_{0}$. But this yields $1>\{(b+k a) \nu\}=\left\{\left(b+k_{0} a\right) \nu\right\}+\left(k-k_{0}\right)\{a \nu\}$ for all $k>k_{0}$, which is absurd.

Similarly, one can prove that $S_{m}$ contains finitely many pairs $(c, d)$ with $c>2 m$, thus finitely many elements at all.

We are now prepared for proving the following crucial lemma.
Lemma. Consider any pair $(a, b)$ with $f(a, b) \neq m$. Then the number $g(a, b)$ of its $m$-excellent descendants is equal to the number $h(a, b)$ of ways to represent the number $t=m-f(a, b)$ as $t=k a+\ell b$ with $k$ and $\ell$ being some nonnegative integers.
Proof. We proceed by induction on the number $N$ of descendants of $(a, b)$ in $S_{m}$. If $N=0$ then clearly $g(a, b)=0$. Assume that $h(a, b)>0$; without loss of generality, we have $a \leqslant b$. Then, clearly, $m-f(a, b) \geqslant a$, so $f(a, b+a) \leqslant f(a, b)+a \leqslant m$ and $a \leqslant m$, hence $(a, b+a) \in S_{m}$ which is impossible. Thus in the base case we have $g(a, b)=h(a, b)=0$, as desired.

Now let $N>0$. Assume that $f(a+b, b)=f(a, b)+b$ and $f(a, b+a)=f(a, b)$ (the other case is similar). If $f(a, b)+b \neq m$, then by the induction hypothesis we have

$$
g(a, b)=g(a+b, b)+g(a, b+a)=h(a+b, b)+h(a, b+a) .
$$

Notice that both pairs $(a+b, b)$ and $(a, b+a)$ are descendants of $(a, b)$ and thus each of them has strictly less descendants in $S_{m}$ than $(a, b)$ does.

Next, each one of the $h(a+b, b)$ representations of $m-f(a+b, b)=m-b-f(a, b)$ as the sum $k^{\prime}(a+b)+\ell^{\prime} b$ provides the representation $m-f(a, b)=k a+\ell b$ with $k=k^{\prime}<k^{\prime}+\ell^{\prime}+1=\ell$. Similarly, each one of the $h(a, b+a)$ representations of $m-f(a, b+a)=m-f(a, b)$ as the sum $k^{\prime} a+\ell^{\prime}(b+a)$ provides the representation $m-f(a, b)=k a+\ell b$ with $k=k^{\prime}+\ell^{\prime} \geqslant \ell^{\prime}=\ell$. This correspondence is obviously bijective, so

$$
h(a, b)=h(a+b, b)+h(a, b+a)=g(a, b),
$$

as required.
Finally, if $f(a, b)+b=m$ then $(a+b, b)$ is $m$-excellent, so $g(a, b)=1+g(a, b+a)=1+h(a, b+a)$ by the induction hypothesis. On the other hand, the number $m-f(a, b)=b$ has a representation $0 \cdot a+1 \cdot b$ and sometimes one more representation as $k a+0 \cdot b$; this last representation exists simultaneously with the representation $m-f(a, b+a)=k a+0 \cdot(b+a)$, so $h(a, b)=1+h(a, b+a)$ as well. Thus in this case the step is also proved.

Now it is easy to finish the solution. There exists a unique $m$-excellent pair of the form $(a, a)$, and each other $m$-excellent pair $(a, b)$ has a unique ancestor of the form $(x, x)$ with $x<m$. By the lemma, for every $x<m$ the number of its $m$-excellent descendants is $h(x, x)$, which is the number of ways to represent $m-f(x, x)=m-x$ as $k x+\ell x$ (with nonnegative integer $k$ and $\ell$ ). This number is 0 if $x \nmid m$, and $m / x$ otherwise. So the total number of excellent pairs is

$$
1+\sum_{x \mid m, x<m} \frac{m}{x}=1+\sum_{d \mid m, d>1} d=\sum_{d \mid m} d,
$$

as required.

Comment. Let us present a sketch of an outline of a different solution. The plan is to check that the number of excellent pairs does not depend on the (irrational) number $\nu$, and to find this number for some appropriate value of $\nu$. For that, we first introduce some geometrical language. We deal only with the excellent pairs ( $a, b$ ) with $a \neq b$.
Part I. Given an irrational positive $\nu$, for every positive integer $n$ we introduce two integral points $F_{\nu}(n)=$ $(n,\lfloor n \nu\rfloor)$ and $C_{\nu}(n)=(n,\lceil n \nu\rceil)$ on the coordinate plane $O x y$. Then (*) reads as $\left[O F_{\nu}(a) C_{\nu}(b)\right]=m / 2$; here [•] stands for the signed area. Next, we rewrite in these terms the condition on a pair $(a, b)$ to be excellent. Let $\ell_{\nu}, \ell_{\nu}^{+}$, and $\ell_{\nu}^{-}$be the lines determined by the equations $y=\nu x, y=\nu x+1$, and $y=\nu x-1$, respectively.
$a)$. Firstly, we deal with all excellent pairs ( $a, b$ ) with $a<b$. Given some value of $a$, all the points $C$ such that $\left[O F_{\nu}(a) C\right]=m / 2$ lie on some line $f_{\nu}(a)$; if there exist any good pairs $(a, b)$ at all, this line has to contain at least one integral point, which happens exactly when $\operatorname{gcd}(a,\lfloor a \nu\rfloor) \mid m$.

Let $P_{\nu}(a)$ be the point of intersection of $\ell_{\nu}^{+}$and $f_{\nu}(a)$, and let $p_{\nu}(a)$ be its abscissa; notice that $p_{\nu}(a)$ is irrational if it is nonzero. Now, if $(a, b)$ is good, then the point $C_{\nu}(b)$ lies on $f_{\nu}(a)$, which means that the point of $f_{\nu}(a)$ with abscissa $b$ lies between $\ell_{\nu}$ and $\ell_{\nu}^{+}$and is integral. If in addition the pair $(a, b-a)$ is not good, then the point of $f_{\nu}(a)$ with abscissa $b-a$ lies above $\ell_{\nu}^{+}$(see Fig. 1). Thus, the pair $(a, b)$ with $b>a$ is excellent exactly when $p_{\nu}(a)$ lies between $b-a$ and $b$, and the point of $f_{\nu}(a)$ with abscissa $b$ is integral (which means that this point is $C_{\nu}(b)$ ).

Notice now that, if $p_{\nu}(a)>a$, then the number of excellent pairs of the form $(a, b)$ (with $\left.b>a\right)$ is $\operatorname{gcd}(a,\lfloor a \nu\rfloor)$.


Figure 1


Figure 2
$b)$. Analogously, considering the pairs $(a, b)$ with $a>b$, we fix the value of $b$, introduce the line $c_{\nu}(b)$ containing all the points $F$ with $\left[O F C_{\nu}(b)\right]=m / 2$, assume that this line contains an integral point (which means $\operatorname{gcd}(b,\lceil b \nu\rceil) \mid m$ ), and denote the common point of $c_{\nu}(b)$ and $\ell_{\nu}^{-}$by $Q_{\nu}(b)$, its abscissa being $q_{\nu}(b)$. Similarly to the previous case, we obtain that the pair $(a, b)$ is excellent exactly when $q_{\nu}(a)$ lies between $a-b$ and $a$, and the point of $c_{\nu}(b)$ with abscissa $a$ is integral (see Fig. 2). Again, if $q_{\nu}(b)>b$, then the number of excellent pairs of the form $(a, b)$ (with $a>b$ ) is $\operatorname{gcd}(b,\lceil b \nu\rceil)$.
Part II, sketchy. Having obtained such a description, one may check how the number of excellent pairs changes as $\nu$ grows. (Having done that, one may find this number for one appropriate value of $\nu$; for instance, it is relatively easy to make this calculation for $\nu \in\left(1,1+\frac{1}{m}\right)$.)

Consider, for the initial value of $\nu$, some excellent pair $(a, t)$ with $a>t$. As $\nu$ grows, this pair eventually stops being excellent; this happens when the point $Q_{\nu}(t)$ passes through $F_{\nu}(a)$. At the same moment, the pair $(a+t, t)$ becomes excellent instead.

This process halts when the point $Q_{\nu}(t)$ eventually disappears, i.e. when $\nu$ passes through the ratio of the coordinates of the point $T=C_{\nu}(t)$. Hence, the point $T$ afterwards is regarded as $F_{\nu}(t)$. Thus, all the old excellent pairs of the form $(a, t)$ with $a>t$ disappear; on the other hand, the same number of excellent pairs with the first element being $t$ just appear.

Similarly, if some pair $(t, b)$ with $t<b$ is initially $\nu$-excellent, then at some moment it stops being excellent when $P_{\nu}(t)$ passes through $C_{\nu}(b)$; at the same moment, the pair $(t, b-t)$ becomes excellent. This process eventually stops when $b-t<t$. At this moment, again the second element of the pair becomes fixed, and the first one starts to increase.

These ideas can be made precise enough to show that the number of excellent pairs remains unchanged, as required.

We should warn the reader that the rigorous elaboration of Part II is technically quite involved, mostly by the reason that the set of moments when the collection of excellent pairs changes is infinite. Especially much care should be applied to the limit points of this set, which are exactly the points when the line $\ell_{\nu}$ passes through some point of the form $C_{\nu}(b)$.

The same ideas may be explained in an algebraic language instead of a geometrical one; the same technicalities remain in this way as well.

## 55th International Mathematical Olympiad

## PROBLEMS SHORT LIST WITH SOLUTIONS



1 MO 2014
Cape Town - South Africa

# Shortlisted Problems with Solutions 

$55^{\text {th }}$ International Mathematical Olympiad
Cape Town, South Africa, 2014

## Note of Confidentiality

## The shortlisted problems should be kept strictly confidential until IMO 2015.

## Contributing Countries

The Organising Committee and the Problem Selection Committee of IMO 2014 thank the following 43 countries for contributing 141 problem proposals.

Australia, Austria, Belgium, Benin, Bulgaria, Colombia, Croatia, Cyprus, Czech Republic, Denmark, Ecuador, Estonia, Finland, France, Georgia, Germany, Greece, Hong Kong, Hungary, Iceland, India, Indonesia, Iran, Ireland, Japan, Lithuania, Luxembourg, Malaysia, Mongolia, Netherlands, Nigeria, Pakistan, Russia, Saudi Arabia, Serbia, Slovakia, Slovenia, South Korea, Thailand, Turkey, Ukraine, United Kingdom, U.S.A.

## Problem Selection Committee

Johan Meyer<br>Ilya I. Bogdanov<br>Géza Kós<br>Waldemar Pompe Christian Reiher<br>Stephan Wagner



## Problems

## Algebra

A1. Let $z_{0}<z_{1}<z_{2}<\cdots$ be an infinite sequence of positive integers. Prove that there exists a unique integer $n \geqslant 1$ such that

$$
z_{n}<\frac{z_{0}+z_{1}+\cdots+z_{n}}{n} \leqslant z_{n+1} .
$$

(Austria)
A2. Define the function $f:(0,1) \rightarrow(0,1)$ by

$$
f(x)= \begin{cases}x+\frac{1}{2} & \text { if } x<\frac{1}{2}, \\ x^{2} & \text { if } x \geqslant \frac{1}{2} .\end{cases}
$$

Let $a$ and $b$ be two real numbers such that $0<a<b<1$. We define the sequences $a_{n}$ and $b_{n}$ by $a_{0}=a, b_{0}=b$, and $a_{n}=f\left(a_{n-1}\right), b_{n}=f\left(b_{n-1}\right)$ for $n>0$. Show that there exists a positive integer $n$ such that

$$
\left(a_{n}-a_{n-1}\right)\left(b_{n}-b_{n-1}\right)<0 .
$$

(Denmark)
A3. For a sequence $x_{1}, x_{2}, \ldots, x_{n}$ of real numbers, we define its price as

$$
\max _{1 \leqslant i \leqslant n}\left|x_{1}+\cdots+x_{i}\right| .
$$

Given $n$ real numbers, Dave and George want to arrange them into a sequence with a low price. Diligent Dave checks all possible ways and finds the minimum possible price $D$. Greedy George, on the other hand, chooses $x_{1}$ such that $\left|x_{1}\right|$ is as small as possible; among the remaining numbers, he chooses $x_{2}$ such that $\left|x_{1}+x_{2}\right|$ is as small as possible, and so on. Thus, in the $i^{\text {th }}$ step he chooses $x_{i}$ among the remaining numbers so as to minimise the value of $\left|x_{1}+x_{2}+\cdots+x_{i}\right|$. In each step, if several numbers provide the same value, George chooses one at random. Finally he gets a sequence with price $G$.

Find the least possible constant $c$ such that for every positive integer $n$, for every collection of $n$ real numbers, and for every possible sequence that George might obtain, the resulting values satisfy the inequality $G \leqslant c D$.
(Georgia)
A4. Determine all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ satisfying

$$
f(f(m)+n)+f(m)=f(n)+f(3 m)+2014
$$

for all integers $m$ and $n$.

A5. Consider all polynomials $P(x)$ with real coefficients that have the following property: for any two real numbers $x$ and $y$ one has

$$
\left|y^{2}-P(x)\right| \leqslant 2|x| \quad \text { if and only if } \quad\left|x^{2}-P(y)\right| \leqslant 2|y|
$$

Determine all possible values of $P(0)$.

A6. Find all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that

$$
n^{2}+4 f(n)=f(f(n))^{2}
$$

for all $n \in \mathbb{Z}$.

## Combinatorics

C1. Let $n$ points be given inside a rectangle $R$ such that no two of them lie on a line parallel to one of the sides of $R$. The rectangle $R$ is to be dissected into smaller rectangles with sides parallel to the sides of $R$ in such a way that none of these rectangles contains any of the given points in its interior. Prove that we have to dissect $R$ into at least $n+1$ smaller rectangles.
(Serbia)
C2. We have $2^{m}$ sheets of paper, with the number 1 written on each of them. We perform the following operation. In every step we choose two distinct sheets; if the numbers on the two sheets are $a$ and $b$, then we erase these numbers and write the number $a+b$ on both sheets. Prove that after $m 2^{m-1}$ steps, the sum of the numbers on all the sheets is at least $4^{m}$.
(Iran)
C3. Let $n \geqslant 2$ be an integer. Consider an $n \times n$ chessboard divided into $n^{2}$ unit squares. We call a configuration of $n$ rooks on this board happy if every row and every column contains exactly one rook. Find the greatest positive integer $k$ such that for every happy configuration of rooks, we can find a $k \times k$ square without a rook on any of its $k^{2}$ unit squares.
(Croatia)
C4. Construct a tetromino by attaching two $2 \times 1$ dominoes along their longer sides such that the midpoint of the longer side of one domino is a corner of the other domino. This construction yields two kinds of tetrominoes with opposite orientations. Let us call them Sand Z-tetrominoes, respectively.


Assume that a lattice polygon $P$ can be tiled with S-tetrominoes. Prove than no matter how we tile $P$ using only S- and Z-tetrominoes, we always use an even number of Z-tetrominoes.
(Hungary)
C5. Consider $n \geqslant 3$ lines in the plane such that no two lines are parallel and no three have a common point. These lines divide the plane into polygonal regions; let $\mathcal{F}$ be the set of regions having finite area. Prove that it is possible to colour $\lceil\sqrt{n / 2}\rceil$ of the lines blue in such a way that no region in $\mathcal{F}$ has a completely blue boundary. (For a real number $x,\lceil x\rceil$ denotes the least integer which is not smaller than $x$.)

C6. We are given an infinite deck of cards, each with a real number on it. For every real number $x$, there is exactly one card in the deck that has $x$ written on it. Now two players draw disjoint sets $A$ and $B$ of 100 cards each from this deck. We would like to define a rule that declares one of them a winner. This rule should satisfy the following conditions:

1. The winner only depends on the relative order of the 200 cards: if the cards are laid down in increasing order face down and we are told which card belongs to which player, but not what numbers are written on them, we can still decide the winner.
2. If we write the elements of both sets in increasing order as $A=\left\{a_{1}, a_{2}, \ldots, a_{100}\right\}$ and $B=\left\{b_{1}, b_{2}, \ldots, b_{100}\right\}$, and $a_{i}>b_{i}$ for all $i$, then $A$ beats $B$.
3. If three players draw three disjoint sets $A, B, C$ from the deck, $A$ beats $B$ and $B$ beats $C$, then $A$ also beats $C$.

How many ways are there to define such a rule? Here, we consider two rules as different if there exist two sets $A$ and $B$ such that $A$ beats $B$ according to one rule, but $B$ beats $A$ according to the other.
(Russia)
C7. Let $M$ be a set of $n \geqslant 4$ points in the plane, no three of which are collinear. Initially these points are connected with $n$ segments so that each point in $M$ is the endpoint of exactly two segments. Then, at each step, one may choose two segments $A B$ and $C D$ sharing a common interior point and replace them by the segments $A C$ and $B D$ if none of them is present at this moment. Prove that it is impossible to perform $n^{3} / 4$ or more such moves.
(Russia)
C8. A card deck consists of 1024 cards. On each card, a set of distinct decimal digits is written in such a way that no two of these sets coincide (thus, one of the cards is empty). Two players alternately take cards from the deck, one card per turn. After the deck is empty, each player checks if he can throw out one of his cards so that each of the ten digits occurs on an even number of his remaining cards. If one player can do this but the other one cannot, the one who can is the winner; otherwise a draw is declared.

Determine all possible first moves of the first player after which he has a winning strategy.
(Russia)
C9. There are $n$ circles drawn on a piece of paper in such a way that any two circles intersect in two points, and no three circles pass through the same point. Turbo the snail slides along the circles in the following fashion. Initially he moves on one of the circles in clockwise direction. Turbo always keeps sliding along the current circle until he reaches an intersection with another circle. Then he continues his journey on this new circle and also changes the direction of moving, i.e. from clockwise to anticlockwise or vice versa.

Suppose that Turbo's path entirely covers all circles. Prove that $n$ must be odd.

## Geometry

G1. The points $P$ and $Q$ are chosen on the side $B C$ of an acute-angled triangle $A B C$ so that $\angle P A B=\angle A C B$ and $\angle Q A C=\angle C B A$. The points $M$ and $N$ are taken on the rays $A P$ and $A Q$, respectively, so that $A P=P M$ and $A Q=Q N$. Prove that the lines $B M$ and $C N$ intersect on the circumcircle of the triangle $A B C$.
(Georgia)
G2. Let $A B C$ be a triangle. The points $K, L$, and $M$ lie on the segments $B C, C A$, and $A B$, respectively, such that the lines $A K, B L$, and $C M$ intersect in a common point. Prove that it is possible to choose two of the triangles $A L M, B M K$, and $C K L$ whose inradii sum up to at least the inradius of the triangle $A B C$.
(Estonia)
G3. Let $\Omega$ and $O$ be the circumcircle and the circumcentre of an acute-angled triangle $A B C$ with $A B>B C$. The angle bisector of $\angle A B C$ intersects $\Omega$ at $M \neq B$. Let $\Gamma$ be the circle with diameter $B M$. The angle bisectors of $\angle A O B$ and $\angle B O C$ intersect $\Gamma$ at points $P$ and $Q$, respectively. The point $R$ is chosen on the line $P Q$ so that $B R=M R$. Prove that $B R \| A C$. (Here we always assume that an angle bisector is a ray.)
(Russia)
G4. Consider a fixed circle $\Gamma$ with three fixed points $A, B$, and $C$ on it. Also, let us fix a real number $\lambda \in(0,1)$. For a variable point $P \notin\{A, B, C\}$ on $\Gamma$, let $M$ be the point on the segment $C P$ such that $C M=\lambda \cdot C P$. Let $Q$ be the second point of intersection of the circumcircles of the triangles $A M P$ and $B M C$. Prove that as $P$ varies, the point $Q$ lies on a fixed circle.
(United Kingdom)
G5. Let $A B C D$ be a convex quadrilateral with $\angle B=\angle D=90^{\circ}$. Point $H$ is the foot of the perpendicular from $A$ to $B D$. The points $S$ and $T$ are chosen on the sides $A B$ and $A D$, respectively, in such a way that $H$ lies inside triangle $S C T$ and

$$
\angle S H C-\angle B S C=90^{\circ}, \quad \angle T H C-\angle D T C=90^{\circ} .
$$

Prove that the circumcircle of triangle $S H T$ is tangent to the line $B D$.
(Iran)
G6. Let $A B C$ be a fixed acute-angled triangle. Consider some points $E$ and $F$ lying on the sides $A C$ and $A B$, respectively, and let $M$ be the midpoint of $E F$. Let the perpendicular bisector of $E F$ intersect the line $B C$ at $K$, and let the perpendicular bisector of $M K$ intersect the lines $A C$ and $A B$ at $S$ and $T$, respectively. We call the pair $(E, F)$ interesting, if the quadrilateral $K S A T$ is cyclic.

Suppose that the pairs $\left(E_{1}, F_{1}\right)$ and $\left(E_{2}, F_{2}\right)$ are interesting. Prove that

$$
\frac{E_{1} E_{2}}{A B}=\frac{F_{1} F_{2}}{A C} .
$$

(Iran)
G7. Let $A B C$ be a triangle with circumcircle $\Omega$ and incentre $I$. Let the line passing through $I$ and perpendicular to $C I$ intersect the segment $B C$ and the $\operatorname{arc} B C$ (not containing $A$ ) of $\Omega$ at points $U$ and $V$, respectively. Let the line passing through $U$ and parallel to $A I$ intersect $A V$ at $X$, and let the line passing through $V$ and parallel to $A I$ intersect $A B$ at $Y$. Let $W$ and $Z$ be the midpoints of $A X$ and $B C$, respectively. Prove that if the points $I, X$, and $Y$ are collinear, then the points $I, W$, and $Z$ are also collinear.
(U.S.A.)

## Number Theory

N1. Let $n \geqslant 2$ be an integer, and let $A_{n}$ be the set

$$
A_{n}=\left\{2^{n}-2^{k} \mid k \in \mathbb{Z}, 0 \leqslant k<n\right\} .
$$

Determine the largest positive integer that cannot be written as the sum of one or more (not necessarily distinct) elements of $A_{n}$.
(Serbia)
N2. Determine all pairs $(x, y)$ of positive integers such that

$$
\begin{equation*}
\sqrt[3]{7 x^{2}-13 x y+7 y^{2}}=|x-y|+1 \tag{U.S.A.}
\end{equation*}
$$

N3. A coin is called a Cape Town coin if its value is $1 / n$ for some positive integer $n$. Given a collection of Cape Town coins of total value at most $99+\frac{1}{2}$, prove that it is possible to split this collection into at most 100 groups each of total value at most 1.
(Luxembourg)
N4. Let $n>1$ be a given integer. Prove that infinitely many terms of the sequence $\left(a_{k}\right)_{k \geqslant 1}$, defined by

$$
a_{k}=\left\lfloor\frac{n^{k}}{k}\right\rfloor
$$

are odd. (For a real number $x,\lfloor x\rfloor$ denotes the largest integer not exceeding $x$.)
(Hong Kong)
N5. Find all triples $(p, x, y)$ consisting of a prime number $p$ and two positive integers $x$ and $y$ such that $x^{p-1}+y$ and $x+y^{p-1}$ are both powers of $p$.
(Belgium)
N6. Let $a_{1}<a_{2}<\cdots<a_{n}$ be pairwise coprime positive integers with $a_{1}$ being prime and $a_{1} \geqslant n+2$. On the segment $I=\left[0, a_{1} a_{2} \cdots a_{n}\right]$ of the real line, mark all integers that are divisible by at least one of the numbers $a_{1}, \ldots, a_{n}$. These points split $I$ into a number of smaller segments. Prove that the sum of the squares of the lengths of these segments is divisible by $a_{1}$.
(Serbia)
N7. Let $c \geqslant 1$ be an integer. Define a sequence of positive integers by $a_{1}=c$ and

$$
a_{n+1}=a_{n}^{3}-4 c \cdot a_{n}^{2}+5 c^{2} \cdot a_{n}+c
$$

for all $n \geqslant 1$. Prove that for each integer $n \geqslant 2$ there exists a prime number $p$ dividing $a_{n}$ but none of the numbers $a_{1}, \ldots, a_{n-1}$.
(Austria)
N8. For every real number $x$, let $\|x\|$ denote the distance between $x$ and the nearest integer. Prove that for every pair $(a, b)$ of positive integers there exist an odd prime $p$ and a positive integer $k$ satisfying

$$
\left\|\frac{a}{p^{k}}\right\|+\left\|\frac{b}{p^{k}}\right\|+\left\|\frac{a+b}{p^{k}}\right\|=1 .
$$

(Hungary)

## Solutions

## Algebra

A1. Let $z_{0}<z_{1}<z_{2}<\cdots$ be an infinite sequence of positive integers. Prove that there exists a unique integer $n \geqslant 1$ such that

$$
\begin{equation*}
z_{n}<\frac{z_{0}+z_{1}+\cdots+z_{n}}{n} \leqslant z_{n+1} . \tag{1}
\end{equation*}
$$

Solution. For $n=1,2, \ldots$ define

$$
d_{n}=\left(z_{0}+z_{1}+\cdots+z_{n}\right)-n z_{n} .
$$

The sign of $d_{n}$ indicates whether the first inequality in (1) holds; i.e., it is satisfied if and only if $d_{n}>0$.

Notice that

$$
n z_{n+1}-\left(z_{0}+z_{1}+\cdots+z_{n}\right)=(n+1) z_{n+1}-\left(z_{0}+z_{1}+\cdots+z_{n}+z_{n+1}\right)=-d_{n+1}
$$

so the second inequality in (1) is equivalent to $d_{n+1} \leqslant 0$. Therefore, we have to prove that there is a unique index $n \geqslant 1$ that satisfies $d_{n}>0 \geqslant d_{n+1}$.

By its definition the sequence $d_{1}, d_{2}, \ldots$ consists of integers and we have

$$
d_{1}=\left(z_{0}+z_{1}\right)-1 \cdot z_{1}=z_{0}>0
$$

From
$d_{n+1}-d_{n}=\left(\left(z_{0}+\cdots+z_{n}+z_{n+1}\right)-(n+1) z_{n+1}\right)-\left(\left(z_{0}+\cdots+z_{n}\right)-n z_{n}\right)=n\left(z_{n}-z_{n+1}\right)<0$
we can see that $d_{n+1}<d_{n}$ and thus the sequence strictly decreases.
Hence, we have a decreasing sequence $d_{1}>d_{2}>\ldots$ of integers such that its first element $d_{1}$ is positive. The sequence must drop below 0 at some point, and thus there is a unique index $n$, that is the index of the last positive term, satisfying $d_{n}>0 \geqslant d_{n+1}$.

Comment. Omitting the assumption that $z_{1}, z_{2}, \ldots$ are integers allows the numbers $d_{n}$ to be all positive. In such cases the desired $n$ does not exist. This happens for example if $z_{n}=2-\frac{1}{2^{n}}$ for all integers $n \geqslant 0$.

A2. Define the function $f:(0,1) \rightarrow(0,1)$ by

$$
f(x)= \begin{cases}x+\frac{1}{2} & \text { if } x<\frac{1}{2}, \\ x^{2} & \text { if } x \geqslant \frac{1}{2} .\end{cases}
$$

Let $a$ and $b$ be two real numbers such that $0<a<b<1$. We define the sequences $a_{n}$ and $b_{n}$ by $a_{0}=a, b_{0}=b$, and $a_{n}=f\left(a_{n-1}\right), b_{n}=f\left(b_{n-1}\right)$ for $n>0$. Show that there exists a positive integer $n$ such that

$$
\left(a_{n}-a_{n-1}\right)\left(b_{n}-b_{n-1}\right)<0 .
$$

(Denmark)
Solution. Note that

$$
f(x)-x=\frac{1}{2}>0
$$

if $x<\frac{1}{2}$ and

$$
f(x)-x=x^{2}-x<0
$$

if $x \geqslant \frac{1}{2}$. So if we consider $(0,1)$ as being divided into the two subintervals $I_{1}=\left(0, \frac{1}{2}\right)$ and $I_{2}=\left[\frac{1}{2}, 1\right)$, the inequality

$$
\left(a_{n}-a_{n-1}\right)\left(b_{n}-b_{n-1}\right)=\left(f\left(a_{n-1}\right)-a_{n-1}\right)\left(f\left(b_{n-1}\right)-b_{n-1}\right)<0
$$

holds if and only if $a_{n-1}$ and $b_{n-1}$ lie in distinct subintervals.
Let us now assume, to the contrary, that $a_{k}$ and $b_{k}$ always lie in the same subinterval. Consider the distance $d_{k}=\left|a_{k}-b_{k}\right|$. If both $a_{k}$ and $b_{k}$ lie in $I_{1}$, then

$$
d_{k+1}=\left|a_{k+1}-b_{k+1}\right|=\left|a_{k}+\frac{1}{2}-b_{k}-\frac{1}{2}\right|=d_{k} .
$$

If, on the other hand, $a_{k}$ and $b_{k}$ both lie in $I_{2}$, then $\min \left(a_{k}, b_{k}\right) \geqslant \frac{1}{2}$ and $\max \left(a_{k}, b_{k}\right)=$ $\min \left(a_{k}, b_{k}\right)+d_{k} \geqslant \frac{1}{2}+d_{k}$, which implies

$$
d_{k+1}=\left|a_{k+1}-b_{k+1}\right|=\left|a_{k}^{2}-b_{k}^{2}\right|=\left|\left(a_{k}-b_{k}\right)\left(a_{k}+b_{k}\right)\right| \geqslant\left|a_{k}-b_{k}\right|\left(\frac{1}{2}+\frac{1}{2}+d_{k}\right)=d_{k}\left(1+d_{k}\right) \geqslant d_{k} .
$$

This means that the difference $d_{k}$ is non-decreasing, and in particular $d_{k} \geqslant d_{0}>0$ for all $k$.
We can even say more. If $a_{k}$ and $b_{k}$ lie in $I_{2}$, then

$$
d_{k+2} \geqslant d_{k+1} \geqslant d_{k}\left(1+d_{k}\right) \geqslant d_{k}\left(1+d_{0}\right)
$$

If $a_{k}$ and $b_{k}$ both lie in $I_{1}$, then $a_{k+1}$ and $b_{k+1}$ both lie in $I_{2}$, and so we have

$$
d_{k+2} \geqslant d_{k+1}\left(1+d_{k+1}\right) \geqslant d_{k+1}\left(1+d_{0}\right)=d_{k}\left(1+d_{0}\right) .
$$

In either case, $d_{k+2} \geqslant d_{k}\left(1+d_{0}\right)$, and inductively we get

$$
d_{2 m} \geqslant d_{0}\left(1+d_{0}\right)^{m} .
$$

For sufficiently large $m$, the right-hand side is greater than 1 , but since $a_{2 m}, b_{2 m}$ both lie in $(0,1)$, we must have $d_{2 m}<1$, a contradiction.

Thus there must be a positive integer $n$ such that $a_{n-1}$ and $b_{n-1}$ do not lie in the same subinterval, which proves the desired statement.

A3. For a sequence $x_{1}, x_{2}, \ldots, x_{n}$ of real numbers, we define its price as

$$
\max _{1 \leqslant i \leqslant n}\left|x_{1}+\cdots+x_{i}\right| .
$$

Given $n$ real numbers, Dave and George want to arrange them into a sequence with a low price. Diligent Dave checks all possible ways and finds the minimum possible price $D$. Greedy George, on the other hand, chooses $x_{1}$ such that $\left|x_{1}\right|$ is as small as possible; among the remaining numbers, he chooses $x_{2}$ such that $\left|x_{1}+x_{2}\right|$ is as small as possible, and so on. Thus, in the $i^{\text {th }}$ step he chooses $x_{i}$ among the remaining numbers so as to minimise the value of $\left|x_{1}+x_{2}+\cdots+x_{i}\right|$. In each step, if several numbers provide the same value, George chooses one at random. Finally he gets a sequence with price $G$.

Find the least possible constant $c$ such that for every positive integer $n$, for every collection of $n$ real numbers, and for every possible sequence that George might obtain, the resulting values satisfy the inequality $G \leqslant c D$.
(Georgia)
Answer. $c=2$.
Solution. If the initial numbers are $1,-1,2$, and -2 , then Dave may arrange them as $1,-2,2,-1$, while George may get the sequence $1,-1,2,-2$, resulting in $D=1$ and $G=2$. So we obtain $c \geqslant 2$.

Therefore, it remains to prove that $G \leqslant 2 D$. Let $x_{1}, x_{2}, \ldots, x_{n}$ be the numbers Dave and George have at their disposal. Assume that Dave and George arrange them into sequences $d_{1}, d_{2}, \ldots, d_{n}$ and $g_{1}, g_{2}, \ldots, g_{n}$, respectively. Put

$$
M=\max _{1 \leqslant i \leqslant n}\left|x_{i}\right|, \quad S=\left|x_{1}+\cdots+x_{n}\right|, \quad \text { and } \quad N=\max \{M, S\}
$$

We claim that

$$
\begin{align*}
& D \geqslant S,  \tag{1}\\
& D \geqslant \frac{M}{2}, \quad \text { and }  \tag{2}\\
& G \leqslant N=\max \{M, S\} . \tag{3}
\end{align*}
$$

These inequalities yield the desired estimate, as $G \leqslant \max \{M, S\} \leqslant \max \{M, 2 S\} \leqslant 2 D$.
The inequality (1) is a direct consequence of the definition of the price.
To prove (2), consider an index $i$ with $\left|d_{i}\right|=M$. Then we have

$$
M=\left|d_{i}\right|=\left|\left(d_{1}+\cdots+d_{i}\right)-\left(d_{1}+\cdots+d_{i-1}\right)\right| \leqslant\left|d_{1}+\cdots+d_{i}\right|+\left|d_{1}+\cdots+d_{i-1}\right| \leqslant 2 D
$$

as required.
It remains to establish (3). Put $h_{i}=g_{1}+g_{2}+\cdots+g_{i}$. We will prove by induction on $i$ that $\left|h_{i}\right| \leqslant N$. The base case $i=1$ holds, since $\left|h_{1}\right|=\left|g_{1}\right| \leqslant M \leqslant N$. Notice also that $\left|h_{n}\right|=S \leqslant N$.

For the induction step, assume that $\left|h_{i-1}\right| \leqslant N$. We distinguish two cases.
Case 1. Assume that no two of the numbers $g_{i}, g_{i+1}, \ldots, g_{n}$ have opposite signs.
Without loss of generality, we may assume that they are all nonnegative. Then one has $h_{i-1} \leqslant h_{i} \leqslant \cdots \leqslant h_{n}$, thus

$$
\left|h_{i}\right| \leqslant \max \left\{\left|h_{i-1}\right|,\left|h_{n}\right|\right\} \leqslant N .
$$

Case 2. Among the numbers $g_{i}, g_{i+1}, \ldots, g_{n}$ there are positive and negative ones.

Then there exists some index $j \geqslant i$ such that $h_{i-1} g_{j} \leqslant 0$. By the definition of George's sequence we have

$$
\left|h_{i}\right|=\left|h_{i-1}+g_{i}\right| \leqslant\left|h_{i-1}+g_{j}\right| \leqslant \max \left\{\left|h_{i-1}\right|,\left|g_{j}\right|\right\} \leqslant N .
$$

Thus, the induction step is established.
Comment 1. One can establish the weaker inequalities $D \geqslant \frac{M}{2}$ and $G \leqslant D+\frac{M}{2}$ from which the result also follows.

Comment 2. One may ask a more specific question to find the maximal suitable $c$ if the number $n$ is fixed. For $n=1$ or 2 , the answer is $c=1$. For $n=3$, the answer is $c=\frac{3}{2}$, and it is reached e.g., for the collection $1,2,-4$. Finally, for $n \geqslant 4$ the answer is $c=2$. In this case the arguments from the solution above apply, and the answer is reached e.g., for the same collection $1,-1,2,-2$, augmented by several zeroes.

A4. Determine all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ satisfying

$$
\begin{equation*}
f(f(m)+n)+f(m)=f(n)+f(3 m)+2014 \tag{1}
\end{equation*}
$$

for all integers $m$ and $n$.
(Netherlands)
Answer. There is only one such function, namely $n \longmapsto 2 n+1007$.
Solution. Let $f$ be a function satisfying (1). Set $C=1007$ and define the function $g: \mathbb{Z} \rightarrow \mathbb{Z}$ by $g(m)=f(3 m)-f(m)+2 C$ for all $m \in \mathbb{Z}$; in particular, $g(0)=2 C$. Now (1) rewrites as

$$
f(f(m)+n)=g(m)+f(n)
$$

for all $m, n \in \mathbb{Z}$. By induction in both directions it follows that

$$
\begin{equation*}
f(t f(m)+n)=t g(m)+f(n) \tag{2}
\end{equation*}
$$

holds for all $m, n, t \in \mathbb{Z}$. Applying this, for any $r \in \mathbb{Z}$, to the triples $(r, 0, f(0))$ and $(0,0, f(r))$ in place of $(m, n, t)$ we obtain

$$
f(0) g(r)=f(f(r) f(0))-f(0)=f(r) g(0)
$$

Now if $f(0)$ vanished, then $g(0)=2 C>0$ would entail that $f$ vanishes identically, contrary to (1). Thus $f(0) \neq 0$ and the previous equation yields $g(r)=\alpha f(r)$, where $\alpha=\frac{g(0)}{f(0)}$ is some nonzero constant.

So the definition of $g$ reveals $f(3 m)=(1+\alpha) f(m)-2 C$, i.e.,

$$
\begin{equation*}
f(3 m)-\beta=(1+\alpha)(f(m)-\beta) \tag{3}
\end{equation*}
$$

for all $m \in \mathbb{Z}$, where $\beta=\frac{2 C}{\alpha}$. By induction on $k$ this implies

$$
\begin{equation*}
f\left(3^{k} m\right)-\beta=(1+\alpha)^{k}(f(m)-\beta) \tag{4}
\end{equation*}
$$

for all integers $k \geqslant 0$ and $m$.
Since $3 \nmid 2014$, there exists by (1) some value $d=f(a)$ attained by $f$ that is not divisible by 3 . Now by (2) we have $f(n+t d)=f(n)+t g(a)=f(n)+\alpha \cdot t f(a)$, i.e.,

$$
\begin{equation*}
f(n+t d)=f(n)+\alpha \cdot t d \tag{5}
\end{equation*}
$$

for all $n, t \in \mathbb{Z}$.
Let us fix any positive integer $k$ with $d \mid\left(3^{k}-1\right)$, which is possible, since $\operatorname{gcd}(3, d)=1$. E.g., by the Euler-Fermat theorem, we may take $k=\varphi(|d|)$. Now for each $m \in \mathbb{Z}$ we get

$$
f\left(3^{k} m\right)=f(m)+\alpha\left(3^{k}-1\right) m
$$

from (5), which in view of (4) yields $\left((1+\alpha)^{k}-1\right)(f(m)-\beta)=\alpha\left(3^{k}-1\right) m$. Since $\alpha \neq 0$, the right hand side does not vanish for $m \neq 0$, wherefore the first factor on the left hand side cannot vanish either. It follows that

$$
f(m)=\frac{\alpha\left(3^{k}-1\right)}{(1+\alpha)^{k}-1} \cdot m+\beta .
$$

So $f$ is a linear function, say $f(m)=A m+\beta$ for all $m \in \mathbb{Z}$ with some constant $A \in \mathbb{Q}$. Plugging this into (1) one obtains $\left(A^{2}-2 A\right) m+(A \beta-2 C)=0$ for all $m$, which is equivalent to the conjunction of

$$
\begin{equation*}
A^{2}=2 A \quad \text { and } \quad A \beta=2 C \tag{6}
\end{equation*}
$$

The first equation is equivalent to $A \in\{0,2\}$, and as $C \neq 0$ the second one gives

$$
\begin{equation*}
A=2 \quad \text { and } \quad \beta=C \tag{7}
\end{equation*}
$$

This shows that $f$ is indeed the function mentioned in the answer and as the numbers found in $(7)$ do indeed satisfy the equations (6) this function is indeed as desired.

Comment 1. One may see that $\alpha=2$. A more pedestrian version of the above solution starts with a direct proof of this fact, that can be obtained by substituting some special values into (1), e.g., as follows.

Set $D=f(0)$. Plugging $m=0$ into (1) and simplifying, we get

$$
\begin{equation*}
f(n+D)=f(n)+2 C \tag{8}
\end{equation*}
$$

for all $n \in \mathbb{Z}$. In particular, for $n=0, D, 2 D$ we obtain $f(D)=2 C+D, f(2 D)=f(D)+2 C=4 C+D$, and $f(3 D)=f(2 D)+2 C=6 C+D$. So substituting $m=D$ and $n=r-D$ into (1) and applying (8) with $n=r-D$ afterwards we learn

$$
f(r+2 C)+2 C+D=(f(r)-2 C)+(6 C+D)+2 C
$$

i.e., $f(r+2 C)=f(r)+4 C$. By induction in both directions it follows that

$$
\begin{equation*}
f(n+2 C t)=f(n)+4 C t \tag{9}
\end{equation*}
$$

holds for all $n, t \in \mathbb{Z}$.
Claim. If $a$ and $b$ denote two integers with the property that $f(n+a)=f(n)+b$ holds for all $n \in \mathbb{Z}$, then $b=2 a$.
Proof. Applying induction in both directions to the assumption we get $f(n+t a)=f(n)+t b$ for all $n, t \in \mathbb{Z}$. Plugging $(n, t)=(0,2 C)$ into this equation and $(n, t)=(0, a)$ into (9) we get $f(2 a C)-f(0)=$ $2 b C=4 a C$, and, as $C \neq 0$, the claim follows.

Now by (1), for any $m \in \mathbb{Z}$, the numbers $a=f(m)$ and $b=f(3 m)-f(m)+2 C$ have the property mentioned in the claim, whence we have

$$
f(3 m)-C=3(f(m)-C)
$$

In view of (3) this tells us indeed that $\alpha=2$.
Now the solution may be completed as above, but due to our knowledge of $\alpha=2$ we get the desired formula $f(m)=2 m+C$ directly without having the need to go through all linear functions. Now it just remains to check that this function does indeed satisfy (1).

Comment 2. It is natural to wonder what happens if one replaces the number 2014 appearing in the statement of the problem by some arbitrary integer $B$.

If $B$ is odd, there is no such function, as can be seen by using the same ideas as in the above solution.

If $B \neq 0$ is even, however, then the only such function is given by $n \longmapsto 2 n+B / 2$. In case $3 \nmid B$ this was essentially proved above, but for the general case one more idea seems to be necessary. Writing $B=3^{\nu} \cdot k$ with some integers $\nu$ and $k$ such that $3 \nmid k$ one can obtain $f(n)=2 n+B / 2$ for all $n$ that are divisible by $3^{\nu}$ in the same manner as usual; then one may use the formula $f(3 n)=3 f(n)-B$ to establish the remaining cases.

Finally, in case $B=0$ there are more solutions than just the function $n \longmapsto 2 n$. It can be shown that all these other functions are periodic; to mention just one kind of example, for any even integers $r$ and $s$ the function

$$
f(n)= \begin{cases}r & \text { if } n \text { is even } \\ s & \text { if } n \text { is odd }\end{cases}
$$

also has the property under discussion.

A5. Consider all polynomials $P(x)$ with real coefficients that have the following property: for any two real numbers $x$ and $y$ one has

$$
\begin{equation*}
\left|y^{2}-P(x)\right| \leqslant 2|x| \quad \text { if and only if } \quad\left|x^{2}-P(y)\right| \leqslant 2|y| . \tag{1}
\end{equation*}
$$

Determine all possible values of $P(0)$.
(Belgium)
Answer. The set of possible values of $P(0)$ is $(-\infty, 0) \cup\{1\}$.

## Solution.

Part I. We begin by verifying that these numbers are indeed possible values of $P(0)$. To see that each negative real number $-C$ can be $P(0)$, it suffices to check that for every $C>0$ the polynomial $P(x)=-\left(\frac{2 x^{2}}{C}+C\right)$ has the property described in the statement of the problem. Due to symmetry it is enough for this purpose to prove $\left|y^{2}-P(x)\right|>2|x|$ for any two real numbers $x$ and $y$. In fact we have

$$
\left|y^{2}-P(x)\right|=y^{2}+\frac{x^{2}}{C}+\frac{(|x|-C)^{2}}{C}+2|x| \geqslant \frac{x^{2}}{C}+2|x| \geqslant 2|x|,
$$

where in the first estimate equality can only hold if $|x|=C$, whilst in the second one it can only hold if $x=0$. As these two conditions cannot be met at the same time, we have indeed $\left|y^{2}-P(x)\right|>2|x|$.

To show that $P(0)=1$ is possible as well, we verify that the polynomial $P(x)=x^{2}+1$ satisfies (1). Notice that for all real numbers $x$ and $y$ we have

$$
\begin{aligned}
\left|y^{2}-P(x)\right| \leqslant 2|x| & \Longleftrightarrow\left(y^{2}-x^{2}-1\right)^{2} \leqslant 4 x^{2} \\
& \Longleftrightarrow 0 \leqslant\left(\left(y^{2}-(x-1)^{2}\right)\left((x+1)^{2}-y^{2}\right)\right. \\
& \Longleftrightarrow 0 \leqslant(y-x+1)(y+x-1)(x+1-y)(x+1+y) \\
& \Longleftrightarrow 0 \leqslant\left((x+y)^{2}-1\right)\left(1-(x-y)^{2}\right) .
\end{aligned}
$$

Since this inequality is symmetric in $x$ and $y$, we are done.
Part II. Now we show that no values other than those mentioned in the answer are possible for $P(0)$. To reach this we let $P$ denote any polynomial satisfying (1) and $P(0) \geqslant 0$; as we shall see, this implies $P(x)=x^{2}+1$ for all real $x$, which is actually more than what we want.

First step: We prove that $P$ is even.
By (1) we have

$$
\left|y^{2}-P(x)\right| \leqslant 2|x| \Longleftrightarrow\left|x^{2}-P(y)\right| \leqslant 2|y| \Longleftrightarrow\left|y^{2}-P(-x)\right| \leqslant 2|x|
$$

for all real numbers $x$ and $y$. Considering just the equivalence of the first and third statement and taking into account that $y^{2}$ may vary through $\mathbb{R}_{\geqslant 0}$ we infer that

$$
[P(x)-2|x|, P(x)+2|x|] \cap \mathbb{R}_{\geqslant 0}=[P(-x)-2|x|, P(-x)+2|x|] \cap \mathbb{R}_{\geqslant 0}
$$

holds for all $x \in \mathbb{R}$. We claim that there are infinitely many real numbers $x$ such that $P(x)+2|x| \geqslant 0$. This holds in fact for any real polynomial with $P(0) \geqslant 0$; in order to see this, we may assume that the coefficient of $P$ appearing in front of $x$ is nonnegative. In this case the desired inequality holds for all sufficiently small positive real numbers.

For such numbers $x$ satisfying $P(x)+2|x| \geqslant 0$ we have $P(x)+2|x|=P(-x)+2|x|$ by the previous displayed formula, and hence also $P(x)=P(-x)$. Consequently the polynomial $P(x)-P(-x)$ has infinitely many zeros, wherefore it has to vanish identically. Thus $P$ is indeed even.

Second step: We prove that $P(t)>0$ for all $t \in \mathbb{R}$.
Let us assume for a moment that there exists a real number $t \neq 0$ with $P(t)=0$. Then there is some open interval $I$ around $t$ such that $|P(y)| \leqslant 2|y|$ holds for all $y \in I$. Plugging $x=0$ into (1) we learn that $y^{2}=P(0)$ holds for all $y \in I$, which is clearly absurd. We have thus shown $P(t) \neq 0$ for all $t \neq 0$.

In combination with $P(0) \geqslant 0$ this informs us that our claim could only fail if $P(0)=0$. In this case there is by our first step a polynomial $Q(x)$ such that $P(x)=x^{2} Q(x)$. Applying (1) to $x=0$ and an arbitrary $y \neq 0$ we get $|y Q(y)|>2$, which is surely false when $y$ is sufficiently small.

Third step: We prove that $P$ is a quadratic polynomial.
Notice that $P$ cannot be constant, for otherwise if $x=\sqrt{P(0)}$ and $y$ is sufficiently large, the first part of (1) is false whilst the second part is true. So the degree $n$ of $P$ has to be at least 1 . By our first step $n$ has to be even as well, whence in particular $n \geqslant 2$.

Now assume that $n \geqslant 4$. Plugging $y=\sqrt{P(x)}$ into (1) we get $\left|x^{2}-P(\sqrt{P(x)})\right| \leqslant 2 \sqrt{P(x)}$ and hence

$$
P(\sqrt{P(x)}) \leqslant x^{2}+2 \sqrt{P(x)}
$$

for all real $x$. Choose positive real numbers $x_{0}, a$, and $b$ such that if $x \in\left(x_{0}, \infty\right)$, then $a x^{n}<$ $P(x)<b x^{n}$; this is indeed possible, for if $d>0$ denotes the leading coefficient of $P$, then $\lim _{x \rightarrow \infty} \frac{P(x)}{x^{n}}=d$, whence for instance the numbers $a=\frac{d}{2}$ and $b=2 d$ work provided that $x_{0}$ is chosen large enough.

Now for all sufficiently large real numbers $x$ we have

$$
a^{n / 2+1} x^{n^{2} / 2}<a P(x)^{n / 2}<P(\sqrt{P(x)}) \leqslant x^{2}+2 \sqrt{P(x)}<x^{n / 2}+2 b^{1 / 2} x^{n / 2}
$$

i.e.

$$
x^{\left(n^{2}-n\right) / 2}<\frac{1+2 b^{1 / 2}}{a^{n / 2+1}}
$$

which is surely absurd. Thus $P$ is indeed a quadratic polynomial.
Fourth step: We prove that $P(x)=x^{2}+1$.
In the light of our first three steps there are two real numbers $a>0$ and $b$ such that $P(x)=$ $a x^{2}+b$. Now if $x$ is large enough and $y=\sqrt{a} x$, the left part of (1) holds and the right part reads $\left|\left(1-a^{2}\right) x^{2}-b\right| \leqslant 2 \sqrt{a} x$. In view of the fact that $a>0$ this is only possible if $a=1$. Finally, substituting $y=x+1$ with $x>0$ into (1) we get

$$
|2 x+1-b| \leqslant 2 x \Longleftrightarrow|2 x+1+b| \leqslant 2 x+2,
$$

i.e.,

$$
b \in[1,4 x+1] \Longleftrightarrow b \in[-4 x-3,1]
$$

for all $x>0$. Choosing $x$ large enough, we can achieve that at least one of these two statements holds; then both hold, which is only possible if $b=1$, as desired.

Comment 1. There are some issues with this problem in that its most natural solutions seem to use some basic facts from analysis, such as the continuity of polynomials or the intermediate value theorem. Yet these facts are intuitively obvious and implicitly clear to the students competing at this level of difficulty, so that the Problem Selection Committee still thinks that the problem is suitable for the IMO.

Comment 2. It seems that most solutions will in the main case, where $P(0)$ is nonnegative, contain an argument that is somewhat asymptotic in nature showing that $P$ is quadratic, and some part narrowing that case down to $P(x)=x^{2}+1$.

Comment 3. It is also possible to skip the first step and start with the second step directly, but then one has to work a bit harder to rule out the case $P(0)=0$. Let us sketch one possibility of doing this: Take the auxiliary polynomial $Q(x)$ such that $P(x)=x Q(x)$. Applying (1) to $x=0$ and an arbitrary $y \neq 0$ we get $|Q(y)|>2$. Hence we either have $Q(z) \geqslant 2$ for all real $z$ or $Q(z) \leqslant-2$ for all real $z$. In particular there is some $\eta \in\{-1,+1\}$ such that $P(\eta) \geqslant 2$ and $P(-\eta) \leqslant-2$. Substituting $x= \pm \eta$ into (1) we learn

$$
\left|y^{2}-P(\eta)\right| \leqslant 2 \Longleftrightarrow|1-P(y)| \leqslant 2|y| \Longleftrightarrow\left|y^{2}-P(-\eta)\right| \leqslant 2 .
$$

But for $y=\sqrt{P(\eta)}$ the first statement is true, whilst the third one is false.
Also, if one has not obtained the evenness of $P$ before embarking on the fourth step, one needs to work a bit harder there, but not in a way that is likely to cause major difficulties.

Comment 4. Truly curious people may wonder about the set of all polynomials having property (1). As explained in the solution above, $P(x)=x^{2}+1$ is the only one with $P(0)=1$. On the other hand, it is not hard to notice that for negative $P(0)$ there are more possibilities than those mentioned above. E.g., as remarked by the proposer, if $a$ and $b$ denote two positive real numbers with $a b>1$ and $Q$ denotes a polynomial attaining nonnegative values only, then $P(x)=-\left(a x^{2}+b+Q(x)\right)$ works.

More generally, it may be proved that if $P(x)$ satisfies (1) and $P(0)<0$, then $-P(x)>2|x|$ holds for all $x \in \mathbb{R}$ so that one just considers the equivalence of two false statements. One may generate all such polynomials $P$ by going through all combinations of a solution of the polynomial equation

$$
x=A(x) B(x)+C(x) D(x)
$$

and a real $E>0$, and setting

$$
P(x)=-\left(A(x)^{2}+B(x)^{2}+C(x)^{2}+D(x)^{2}+E\right)
$$

for each of them.

A6. Find all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that

$$
\begin{equation*}
n^{2}+4 f(n)=f(f(n))^{2} \tag{1}
\end{equation*}
$$

for all $n \in \mathbb{Z}$.
(United Kingdom)
Answer. The possibilities are:

- $f(n)=n+1$ for all $n$;
- or, for some $a \geqslant 1, \quad f(n)= \begin{cases}n+1, & n>-a, \\ -n+1, & n \leqslant-a ;\end{cases}$
- or $f(n)= \begin{cases}n+1, & n>0, \\ 0, & n=0, \\ -n+1, & n<0 .\end{cases}$


## Solution 1.

Part I. Let us first check that each of the functions above really satisfies the given functional equation. If $f(n)=n+1$ for all $n$, then we have

$$
n^{2}+4 f(n)=n^{2}+4 n+4=(n+2)^{2}=f(n+1)^{2}=f(f(n))^{2} .
$$

If $f(n)=n+1$ for $n>-a$ and $f(n)=-n+1$ otherwise, then we have the same identity for $n>-a$ and

$$
n^{2}+4 f(n)=n^{2}-4 n+4=(2-n)^{2}=f(1-n)^{2}=f(f(n))^{2}
$$

otherwise. The same applies to the third solution (with $a=0$ ), where in addition one has

$$
0^{2}+4 f(0)=0=f(f(0))^{2}
$$

Part II. It remains to prove that these are really the only functions that satisfy our functional equation. We do so in three steps:

Step 1: We prove that $f(n)=n+1$ for $n>0$.
Consider the sequence $\left(a_{k}\right)$ given by $a_{k}=f^{k}(1)$ for $k \geqslant 0$. Setting $n=a_{k}$ in (1), we get

$$
a_{k}^{2}+4 a_{k+1}=a_{k+2}^{2}
$$

Of course, $a_{0}=1$ by definition. Since $a_{2}^{2}=1+4 a_{1}$ is odd, $a_{2}$ has to be odd as well, so we set $a_{2}=2 r+1$ for some $r \in \mathbb{Z}$. Then $a_{1}=r^{2}+r$ and consequently

$$
a_{3}^{2}=a_{1}^{2}+4 a_{2}=\left(r^{2}+r\right)^{2}+8 r+4
$$

Since $8 r+4 \neq 0, a_{3}^{2} \neq\left(r^{2}+r\right)^{2}$, so the difference between $a_{3}^{2}$ and $\left(r^{2}+r\right)^{2}$ is at least the distance from $\left(r^{2}+r\right)^{2}$ to the nearest even square (since $8 r+4$ and $r^{2}+r$ are both even). This implies that

$$
|8 r+4|=\left|a_{3}^{2}-\left(r^{2}+r\right)^{2}\right| \geqslant\left(r^{2}+r\right)^{2}-\left(r^{2}+r-2\right)^{2}=4\left(r^{2}+r-1\right)
$$

(for $r=0$ and $r=-1$, the estimate is trivial, but this does not matter). Therefore, we ave

$$
4 r^{2} \leqslant|8 r+4|-4 r+4
$$

If $|r| \geqslant 4$, then

$$
4 r^{2} \geqslant 16|r| \geqslant 12|r|+16>8|r|+4+4|r|+4 \geqslant|8 r+4|-4 r+4,
$$

a contradiction. Thus $|r|<4$. Checking all possible remaining values of $r$, we find that $\left(r^{2}+r\right)^{2}+8 r+4$ is only a square in three cases: $r=-3, r=0$ and $r=1$. Let us now distinguish these three cases:

- $r=-3$, thus $a_{1}=6$ and $a_{2}=-5$. For each $k \geqslant 1$, we have

$$
a_{k+2}= \pm \sqrt{a_{k}^{2}+4 a_{k+1}}
$$

and the sign needs to be chosen in such a way that $a_{k+1}^{2}+4 a_{k+2}$ is again a square. This yields $a_{3}=-4, a_{4}=-3, a_{5}=-2, a_{6}=-1, a_{7}=0, a_{8}=1, a_{9}=2$. At this point we have reached a contradiction, since $f(1)=f\left(a_{0}\right)=a_{1}=6$ and at the same time $f(1)=f\left(a_{8}\right)=a_{9}=2$.

- $r=0$, thus $a_{1}=0$ and $a_{2}=1$. Then $a_{3}^{2}=a_{1}^{2}+4 a_{2}=4$, so $a_{3}= \pm 2$. This, however, is a contradiction again, since it gives us $f(1)=f\left(a_{0}\right)=a_{1}=0$ and at the same time $f(1)=f\left(a_{2}\right)=a_{3}= \pm 2$.
- $r=1$, thus $a_{1}=2$ and $a_{2}=3$. We prove by induction that $a_{k}=k+1$ for all $k \geqslant 0$ in this case, which we already know for $k \leqslant 2$ now. For the induction step, assume that $a_{k-1}=k$ and $a_{k}=k+1$. Then

$$
a_{k+1}^{2}=a_{k-1}^{2}+4 a_{k}=k^{2}+4 k+4=(k+2)^{2}
$$

so $a_{k+1}= \pm(k+2)$. If $a_{k+1}=-(k+2)$, then

$$
a_{k+2}^{2}=a_{k}^{2}+4 a_{k+1}=(k+1)^{2}-4 k-8=k^{2}-2 k-7=(k-1)^{2}-8 .
$$

The latter can only be a square if $k=4$ (since 1 and 9 are the only two squares whose difference is 8 ). Then, however, $a_{4}=5, a_{5}=-6$ and $a_{6}= \pm 1$, so

$$
a_{7}^{2}=a_{5}^{2}+4 a_{6}=36 \pm 4,
$$

but neither 32 nor 40 is a perfect square. Thus $a_{k+1}=k+2$, which completes our induction. This also means that $f(n)=f\left(a_{n-1}\right)=a_{n}=n+1$ for all $n \geqslant 1$.

Step 2: We prove that either $f(0)=1$, or $f(0)=0$ and $f(n) \neq 0$ for $n \neq 0$.
Set $n=0$ in (1) to get

$$
4 f(0)=f(f(0))^{2} .
$$

This means that $f(0) \geqslant 0$. If $f(0)=0$, then $f(n) \neq 0$ for all $n \neq 0$, since we would otherwise have

$$
n^{2}=n^{2}+4 f(n)=f(f(n))^{2}=f(0)^{2}=0
$$

If $f(0)>0$, then we know that $f(f(0))=f(0)+1$ from the first step, so

$$
4 f(0)=(f(0)+1)^{2}
$$

which yields $f(0)=1$.

Step 3: We discuss the values of $f(n)$ for $n<0$.
Lemma. For every $n \geqslant 1$, we have $f(-n)=-n+1$ or $f(-n)=n+1$. Moreover, if $f(-n)=$ $-n+1$ for some $n \geqslant 1$, then also $f(-n+1)=-n+2$.
Proof. We prove this statement by strong induction on $n$. For $n=1$, we get

$$
1+4 f(-1)=f(f(-1))^{2}
$$

Thus $f(-1)$ needs to be nonnegative. If $f(-1)=0$, then $f(f(-1))=f(0)= \pm 1$, so $f(0)=1$ (by our second step). Otherwise, we know that $f(f(-1))=f(-1)+1$, so

$$
1+4 f(-1)=(f(-1)+1)^{2}
$$

which yields $f(-1)=2$ and thus establishes the base case. For the induction step, we consider two cases:

- If $f(-n) \leqslant-n$, then

$$
f(f(-n))^{2}=(-n)^{2}+4 f(-n) \leqslant n^{2}-4 n<(n-2)^{2}
$$

so $|f(f(-n))| \leqslant n-3$ (for $n=2$, this case cannot even occur). If $f(f(-n)) \geqslant 0$, then we already know from the first two steps that $f(f(f(-n)))=f(f(-n))+1$, unless perhaps if $f(0)=0$ and $f(f(-n))=0$. However, the latter would imply $f(-n)=0$ (as shown in Step 2) and thus $n=0$, which is impossible. If $f(f(-n))<0$, we can apply the induction hypothesis to $f(f(-n))$. In either case, $f(f(f(-n)))= \pm f(f(-n))+1$. Therefore,

$$
f(-n)^{2}+4 f(f(-n))=f(f(f(-n)))^{2}=( \pm f(f(-n))+1)^{2}
$$

which gives us

$$
\begin{aligned}
n^{2} & \leqslant f(-n)^{2}=( \pm f(f(-n))+1)^{2}-4 f(f(-n)) \leqslant f(f(-n))^{2}+6|f(f(-n))|+1 \\
& \leqslant(n-3)^{2}+6(n-3)+1=n^{2}-8
\end{aligned}
$$

a contradiction.

- Thus, we are left with the case that $f(-n)>-n$. Now we argue as in the previous case: if $f(-n) \geqslant 0$, then $f(f(-n))=f(-n)+1$ by the first two steps, since $f(0)=0$ and $f(-n)=0$ would imply $n=0$ (as seen in Step 2) and is thus impossible. If $f(-n)<0$, we can apply the induction hypothesis, so in any case we can infer that $f(f(-n))= \pm f(-n)+1$. We obtain

$$
(-n)^{2}+4 f(-n)=( \pm f(-n)+1)^{2}
$$

so either

$$
n^{2}=f(-n)^{2}-2 f(-n)+1=(f(-n)-1)^{2}
$$

which gives us $f(-n)= \pm n+1$, or

$$
n^{2}=f(-n)^{2}-6 f(-n)+1=(f(-n)-3)^{2}-8 .
$$

Since 1 and 9 are the only perfect squares whose difference is 8 , we must have $n=1$, which we have already considered.

Finally, suppose that $f(-n)=-n+1$ for some $n \geqslant 2$. Then

$$
f(-n+1)^{2}=f(f(-n))^{2}=(-n)^{2}+4 f(-n)=(n-2)^{2}
$$

so $f(-n+1)= \pm(n-2)$. However, we already know that $f(-n+1)=-n+2$ or $f(-n+1)=n$, so $f(-n+1)=-n+2$.

Combining everything we know, we find the solutions as stated in the answer:

- One solution is given by $f(n)=n+1$ for all $n$.
- If $f(n)$ is not always equal to $n+1$, then there is a largest integer $m$ (which cannot be positive) for which this is not the case. In view of the lemma that we proved, we must then have $f(n)=-n+1$ for any integer $n<m$. If $m=-a<0$, we obtain $f(n)=-n+1$ for $n \leqslant-a$ (and $f(n)=n+1$ otherwise). If $m=0$, we have the additional possibility that $f(0)=0, f(n)=-n+1$ for negative $n$ and $f(n)=n+1$ for positive $n$.

Solution 2. Let us provide an alternative proof for Part II, which also proceeds in several steps.

Step 1. Let $a$ be an arbitrary integer and $b=f(a)$. We first concentrate on the case where $|a|$ is sufficiently large.

1. If $b=0$, then (1) applied to $a$ yields $a^{2}=f(f(a))^{2}$, thus

$$
\begin{equation*}
f(a)=0 \quad \Rightarrow \quad a= \pm f(0) . \tag{2}
\end{equation*}
$$

From now on, we set $D=|f(0)|$. Throughout Step 1, we will assume that $a \notin\{-D, 0, D\}$, thus $b \neq 0$.
2. From (1), noticing that $f(f(a))$ and $a$ have the same parity, we get

$$
0 \neq 4|b|=\left|f(f(a))^{2}-a^{2}\right| \geqslant a^{2}-(|a|-2)^{2}=4|a|-4 .
$$

Hence we have

$$
\begin{equation*}
|b|=|f(a)| \geqslant|a|-1 \quad \text { for } a \notin\{-D, 0, D\} . \tag{3}
\end{equation*}
$$

For the rest of Step 1, we also assume that $|a| \geqslant E=\max \{D+2,10\}$. Then by (3) we have $|b| \geqslant D+1$ and thus $|f(b)| \geqslant D$.
3. Set $c=f(b)$; by (3), we have $|c| \geqslant|b|-1$. Thus (1) yields

$$
a^{2}+4 b=c^{2} \geqslant(|b|-1)^{2},
$$

which implies

$$
a^{2} \geqslant(|b|-1)^{2}-4|b|=(|b|-3)^{2}-8>(|b|-4)^{2}
$$

because $|b| \geqslant|a|-1 \geqslant 9$. Thus (3) can be refined to

$$
|a|+3 \geqslant|f(a)| \geqslant|a|-1 \quad \text { for }|a| \geqslant E
$$

Now, from $c^{2}=a^{2}+4 b$ with $|b| \in[|a|-1,|a|+3]$ we get $c^{2}=(a \pm 2)^{2}+d$, where $d \in\{-16,-12,-8,-4,0,4,8\}$. Since $|a \pm 2| \geqslant 8$, this can happen only if $c^{2}=(a \pm 2)^{2}$, which in turn yields $b= \pm a+1$. To summarise,

$$
\begin{equation*}
f(a)=1 \pm a \quad \text { for }|a| \geqslant E . \tag{4}
\end{equation*}
$$

We have shown that, with at most finitely many exceptions, $f(a)=1 \pm a$. Thus it will be convenient for our second step to introduce the sets

$$
Z_{+}=\{a \in \mathbb{Z}: f(a)=a+1\}, \quad Z_{-}=\{a \in \mathbb{Z}: f(a)=1-a\}, \quad \text { and } \quad Z_{0}=\mathbb{Z} \backslash\left(Z_{+} \cup Z_{-}\right)
$$

Step 2. Now we investigate the structure of the sets $Z_{+}, Z_{-}$, and $Z_{0}$.
4. Note that $f(E+1)=1 \pm(E+1)$. If $f(E+1)=E+2$, then $E+1 \in Z_{+}$. Otherwise we have $f(1+E)=-E$; then the original equation (1) with $n=E+1$ gives us $(E-1)^{2}=f(-E)^{2}$, so $f(-E)= \pm(E-1)$. By (4) this may happen only if $f(-E)=1-E$, so in this case $-E \in Z_{+}$. In any case we find that $Z_{+} \neq \varnothing$.
5. Now take any $a \in Z_{+}$. We claim that every integer $x \geqslant a$ also lies in $Z_{+}$. We proceed by induction on $x$, the base case $x=a$ being covered by our assumption. For the induction step, assume that $f(x-1)=x$ and plug $n=x-1$ into (1). We get $f(x)^{2}=(x+1)^{2}$, so either $f(x)=x+1$ or $f(x)=-(x+1)$.
Assume that $f(x)=-(x+1)$ and $x \neq-1$, since otherwise we already have $f(x)=x+1$. Plugging $n=x$ into (1), we obtain $f(-x-1)^{2}=(x-2)^{2}-8$, which may happen only if $x-2= \pm 3$ and $f(-x-1)= \pm 1$. Plugging $n=-x-1$ into (1), we get $f( \pm 1)^{2}=(x+1)^{2} \pm 4$, which in turn may happen only if $x+1 \in\{-2,0,2\}$.
Thus $x \in\{-1,5\}$ and at the same time $x \in\{-3,-1,1\}$, which gives us $x=-1$. Since this has already been excluded, we must have $f(x)=x+1$, which completes our induction.
6. Now we know that either $Z_{+}=\mathbb{Z}$ (if $Z_{+}$is not bounded below), or $Z_{+}=\left\{a \in \mathbb{Z}: a \geqslant a_{0}\right\}$, where $a_{0}$ is the smallest element of $Z_{+}$. In the former case, $f(n)=n+1$ for all $n \in \mathbb{Z}$, which is our first solution. So we assume in the following that $Z_{+}$is bounded below and has a smallest element $a_{0}$.
If $Z_{0}=\varnothing$, then we have $f(x)=x+1$ for $x \geqslant a_{0}$ and $f(x)=1-x$ for $x<a_{0}$. In particular, $f(0)=1$ in any case, so $0 \in Z_{+}$and thus $a_{0} \leqslant 0$. Thus we end up with the second solution listed in the answer. It remains to consider the case where $Z_{0} \neq \varnothing$.
7. Assume that there exists some $a \in Z_{0}$ with $b=f(a) \notin Z_{0}$, so that $f(b)=1 \pm b$. Then we have $a^{2}+4 b=(1 \pm b)^{2}$, so either $a^{2}=(b-1)^{2}$ or $a^{2}=(b-3)^{2}-8$. In the former case we have $b=1 \pm a$, which is impossible by our choice of $a$. So we get $a^{2}=(b-3)^{2}-8$, which implies $f(b)=1-b$ and $|a|=1,|b-3|=3$.
If $b=0$, then we have $f(b)=1$, so $b \in Z_{+}$and therefore $a_{0} \leqslant 0$; hence $a=-1$. But then $f(a)=0=a+1$, so $a \in Z_{+}$, which is impossible.
If $b=6$, then we have $f(6)=-5$, so $f(-5)^{2}=16$ and $f(-5) \in\{-4,4\}$. Then $f(f(-5))^{2}=$ $25+4 f(-5) \in\{9,41\}$, so $f(-5)=-4$ and $-5 \in Z_{+}$. This implies $a_{0} \leqslant-5$, which contradicts our assumption that $\pm 1=a \notin Z_{+}$.
8. Thus we have shown that $f\left(Z_{0}\right) \subseteq Z_{0}$, and $Z_{0}$ is finite. Take any element $c \in Z_{0}$, and consider the sequence defined by $c_{i}=f^{i}(c)$. All elements of the sequence $\left(c_{i}\right)$ lie in $Z_{0}$, hence it is bounded. Choose an index $k$ for which $\left|c_{k}\right|$ is maximal, so that in particular $\left|c_{k+1}\right| \leqslant\left|c_{k}\right|$ and $\left|c_{k+2}\right| \leqslant\left|c_{k}\right|$. Our functional equation (1) yields

$$
\left(\left|c_{k}\right|-2\right)^{2}-4=\left|c_{k}\right|^{2}-4\left|c_{k}\right| \leqslant c_{k}^{2}+4 c_{k+1}=c_{k+2}^{2}
$$

Since $c_{k}$ and $c_{k+2}$ have the same parity and $\left|c_{k+2}\right| \leqslant\left|c_{k}\right|$, this leaves us with three possibilities: $\left|c_{k+2}\right|=\left|c_{k}\right|,\left|c_{k+2}\right|=\left|c_{k}\right|-2$, and $\left|c_{k}\right|-2= \pm 2, c_{k+2}=0$.

If $\left|c_{k+2}\right|=\left|c_{k}\right|-2$, then $f\left(c_{k}\right)=c_{k+1}=1-\left|c_{k}\right|$, which means that $c_{k} \in Z_{-}$or $c_{k} \in Z_{+}$, and we reach a contradiction.

If $\left|c_{k+2}\right|=\left|c_{k}\right|$, then $c_{k+1}=0$, thus $c_{k+3}^{2}=4 c_{k+2}$. So either $c_{k+3} \neq 0$ or (by maximality of $\left.\left|c_{k+2}\right|=\left|c_{k}\right|\right) c_{i}=0$ for all $i$. In the former case, we can repeat the entire argument
with $c_{k+2}$ in the place of $c_{k}$. Now $\left|c_{k+4}\right|=\left|c_{k+2}\right|$ is not possible any more since $c_{k+3} \neq 0$, leaving us with the only possibility $\left|c_{k}\right|-2=\left|c_{k+2}\right|-2= \pm 2$.

Thus we know now that either all $c_{i}$ are equal to 0 , or $\left|c_{k}\right|=4$. If $c_{k}= \pm 4$, then either $c_{k+1}=0$ and $\left|c_{k+2}\right|=\left|c_{k}\right|=4$, or $c_{k+2}=0$ and $c_{k+1}=-4$. From this point onwards, all elements of the sequence are either 0 or $\pm 4$.
Let $c_{r}$ be the last element of the sequence that is not equal to 0 or $\pm 4$ (if such an element exists). Then $c_{r+1}, c_{r+2} \in\{-4,0,4\}$, so

$$
c_{r}^{2}=c_{r+2}^{2}-4 c_{r+1} \in\{-16,0,16,32\}
$$

which gives us a contradiction. Thus all elements of the sequence are equal to 0 or $\pm 4$, and since the choice of $c_{0}=c$ was arbitrary, $Z_{0} \subseteq\{-4,0,4\}$.
9. Finally, we show that $4 \notin Z_{0}$ and $-4 \notin Z_{0}$. Suppose that $4 \in Z_{0}$. Then in particular $a_{0}$ (the smallest element of $Z_{+}$) cannot be less than 4 , since this would imply $4 \in Z_{+}$. So $-3 \in Z_{-}$, which means that $f(-3)=4$. Then $25=(-3)^{2}+4 f(-3)=f(f(-3))^{2}=f(4)^{2}$, so $f(4)= \pm 5 \notin Z_{0}$, and we reach a contradiction.
Suppose that $-4 \in Z_{0}$. The only possible values for $f(-4)$ that are left are 0 and -4 . Note that $4 f(0)=f(f(0))^{2}$, so $f(0) \geqslant 0$. If $f(-4)=0$, then we get $16=(-4)^{2}+0=f(0)^{2}$, thus $f(0)=4$. But then $f(f(-4)) \notin Z_{0}$, which is impossible. Thus $f(-4)=-4$, which gives us $0=(-4)^{2}+4 f(-4)=f(f(-4))^{2}=16$, and this is clearly absurd.
Now we are left with $Z_{0}=\{0\}$ and $f(0)=0$ as the only possibility. If $1 \in Z_{-}$, then $f(1)=0$, so $1=1^{2}+4 f(1)=f(f(1))^{2}=f(0)^{2}=0$, which is another contradiction. Thus $1 \in Z_{+}$, meaning that $a_{0} \leqslant 1$. On the other hand, $a_{0} \leqslant 0$ would imply $0 \in Z_{+}$, so we can only have $a_{0}=1$. Thus $Z_{+}$comprises all positive integers, and $Z_{-}$comprises all negative integers. This gives us the third solution.

Comment. All solutions known to the Problem Selection Committee are quite lengthy and technical, as the two solutions presented here show. It is possible to make the problem easier by imposing additional assumptions, such as $f(0) \neq 0$ or $f(n) \geqslant 1$ for all $n \geqslant 0$, to remove some of the technicalities.

## Combinatorics

C1. Let $n$ points be given inside a rectangle $R$ such that no two of them lie on a line parallel to one of the sides of $R$. The rectangle $R$ is to be dissected into smaller rectangles with sides parallel to the sides of $R$ in such a way that none of these rectangles contains any of the given points in its interior. Prove that we have to dissect $R$ into at least $n+1$ smaller rectangles.
(Serbia)
Solution 1. Let $k$ be the number of rectangles in the dissection. The set of all points that are corners of one of the rectangles can be divided into three disjoint subsets:

- $A$, which consists of the four corners of the original rectangle $R$, each of which is the corner of exactly one of the smaller rectangles,
- $B$, which contains points where exactly two of the rectangles have a common corner (T-junctions, see the figure below),
- $C$, which contains points where four of the rectangles have a common corner (crossings, see the figure below).


Figure 1: A T-junction and a crossing
We denote the number of points in $B$ by $b$ and the number of points in $C$ by $c$. Since each of the $k$ rectangles has exactly four corners, we get

$$
4 k=4+2 b+4 c .
$$

It follows that $2 b \leqslant 4 k-4$, so $b \leqslant 2 k-2$.
Each of the $n$ given points has to lie on a side of one of the smaller rectangles (but not of the original rectangle $R$ ). If we extend this side as far as possible along borders between rectangles, we obtain a line segment whose ends are T-junctions. Note that every point in $B$ can only be an endpoint of at most one such segment containing one of the given points, since it is stated that no two of them lie on a common line parallel to the sides of $R$. This means that

$$
b \geqslant 2 n .
$$

Combining our two inequalities for $b$, we get

$$
2 k-2 \geqslant b \geqslant 2 n,
$$

thus $k \geqslant n+1$, which is what we wanted to prove.

Solution 2. Let $k$ denote the number of rectangles. In the following, we refer to the directions of the sides of $R$ as 'horizontal' and 'vertical' respectively. Our goal is to prove the inequality $k \geqslant n+1$ for fixed $n$. Equivalently, we can prove the inequality $n \leqslant k-1$ for each $k$, which will be done by induction on $k$. For $k=1$, the statement is trivial.

Now assume that $k>1$. If none of the line segments that form the borders between the rectangles is horizontal, then we have $k-1$ vertical segments dividing $R$ into $k$ rectangles. On each of them, there can only be one of the $n$ points, so $n \leqslant k-1$, which is exactly what we want to prove.

Otherwise, consider the lowest horizontal line $h$ that contains one or more of these line segments. Let $R^{\prime}$ be the rectangle that results when everything that lies below $h$ is removed from $R$ (see the example in the figure below).

The rectangles that lie entirely below $h$ form blocks of rectangles separated by vertical line segments. Suppose there are $r$ blocks and $k_{i}$ rectangles in the $i^{\text {th }}$ block. The left and right border of each block has to extend further upwards beyond $h$. Thus we can move any points that lie on these borders upwards, so that they now lie inside $R^{\prime}$. This can be done without violating the conditions, one only needs to make sure that they do not get to lie on a common horizontal line with one of the other given points.

All other borders between rectangles in the $i^{\text {th }}$ block have to lie entirely below $h$. There are $k_{i}-1$ such line segments, each of which can contain at most one of the given points. Finally, there can be one point that lies on $h$. All other points have to lie in $R^{\prime}$ (after moving some of them as explained in the previous paragraph).


Figure 2: Illustration of the inductive argument
We see that $R^{\prime}$ is divided into $k-\sum_{i=1}^{r} k_{i}$ rectangles. Applying the induction hypothesis to $R^{\prime}$, we find that there are at most

$$
\left(k-\sum_{i=1}^{r} k_{i}\right)-1+\sum_{i=1}^{r}\left(k_{i}-1\right)+1=k-r
$$

points. Since $r \geqslant 1$, this means that $n \leqslant k-1$, which completes our induction.

C2. We have $2^{m}$ sheets of paper, with the number 1 written on each of them. We perform the following operation. In every step we choose two distinct sheets; if the numbers on the two sheets are $a$ and $b$, then we erase these numbers and write the number $a+b$ on both sheets. Prove that after $m 2^{m-1}$ steps, the sum of the numbers on all the sheets is at least $4^{m}$.
(Iran)
Solution. Let $P_{k}$ be the product of the numbers on the sheets after $k$ steps.
Suppose that in the $(k+1)^{\text {th }}$ step the numbers $a$ and $b$ are replaced by $a+b$. In the product, the number $a b$ is replaced by $(a+b)^{2}$, and the other factors do not change. Since $(a+b)^{2} \geqslant 4 a b$, we see that $P_{k+1} \geqslant 4 P_{k}$. Starting with $P_{0}=1$, a straightforward induction yields

$$
P_{k} \geqslant 4^{k}
$$

for all integers $k \geqslant 0$; in particular

$$
P_{m \cdot 2^{m-1}} \geqslant 4^{m \cdot 2^{m-1}}=\left(2^{m}\right)^{2^{m}}
$$

so by the AM-GM inequality, the sum of the numbers written on the sheets after $m 2^{m-1}$ steps is at least

$$
2^{m} \cdot \sqrt[2^{m}]{P_{m \cdot 2^{m-1}}} \geqslant 2^{m} \cdot 2^{m}=4^{m}
$$

Comment 1. It is possible to achieve the sum $4^{m}$ in $m 2^{m-1}$ steps. For example, starting from $2^{m}$ equal numbers on the sheets, in $2^{m-1}$ consecutive steps we can double all numbers. After $m$ such doubling rounds we have the number $2^{m}$ on every sheet.

Comment 2. There are several versions of the solution above. E.g., one may try to assign to each positive integer $n$ a weight $w_{n}$ in such a way that the sum of the weights of the numbers written on the sheets increases, say, by at least 2 in each step. For this purpose, one needs the inequality

$$
\begin{equation*}
2 w_{a+b} \geqslant w_{a}+w_{b}+2 \tag{1}
\end{equation*}
$$

to be satisfied for all positive integers $a$ and $b$.
Starting from $w_{1}=1$ and trying to choose the weights as small as possible, one may find that these weights can be defined as follows: For every positive integer $n$, one chooses $k$ to be the maximal integer such that $n \geqslant 2^{k}$, and puts

$$
\begin{equation*}
w_{n}=k+\frac{n}{2^{k}}=\min _{d \in \mathbb{Z} \geqslant 0}\left(d+\frac{n}{2^{d}}\right) . \tag{2}
\end{equation*}
$$

Now, in order to prove that these weights satisfy (1), one may take arbitrary positive integers $a$ and $b$, and choose an integer $d \geqslant 0$ such that $w_{a+b}=d+\frac{a+b}{2^{d}}$. Then one has

$$
2 w_{a+b}=2 d+2 \cdot \frac{a+b}{2^{d}}=\left((d-1)+\frac{a}{2^{d-1}}\right)+\left((d-1)+\frac{b}{2^{d-1}}\right)+2 \geqslant w_{a}+w_{b}+2
$$

Since the initial sum of the weights was $2^{m}$, after $m 2^{m-1}$ steps the sum is at least $(m+1) 2^{m}$. To finish the solution, one may notice that by (2) for every positive integer $a$ one has

$$
\begin{equation*}
w_{a} \leqslant m+\frac{a}{2^{m}}, \quad \text { i.e., } \quad a \geqslant 2^{m}\left(-m+w_{a}\right) \tag{3}
\end{equation*}
$$

So the sum of the numbers $a_{1}, a_{2}, \ldots, a_{2^{m}}$ on the sheets can be estimated as

$$
\sum_{i=1}^{2^{m}} a_{i} \geqslant \sum_{i=1}^{2^{m}} 2^{m}\left(-m+w_{a_{i}}\right)=-m 2^{m} \cdot 2^{m}+2^{m} \sum_{i=1}^{2^{m}} w_{a_{i}} \geqslant-m 4^{m}+(m+1) 4^{m}=4^{m}
$$

as required.
For establishing the inequalities (1) and (3), one may also use the convexity argument, instead of the second definition of $w_{n}$ in (2).

One may check that $\log _{2} n \leqslant w_{n} \leqslant \log _{2} n+1$; thus, in some rough sense, this approach is obtained by "taking the logarithm" of the solution above.

Comment 3. An intuitive strategy to minimise the sum of numbers is that in every step we choose the two smallest numbers. We may call this the greedy strategy. In the following paragraphs we prove that the greedy strategy indeed provides the least possible sum of numbers.

Claim. Starting from any sequence $x_{1}, \ldots, x_{N}$ of positive real numbers on $N$ sheets, for any number $k$ of steps, the greedy strategy achieves the lowest possible sum of numbers.

Proof. We apply induction on $k$; for $k=1$ the statement is obvious. Let $k \geqslant 2$, and assume that the claim is true for smaller values.

Every sequence of $k$ steps can be encoded as $S=\left(\left(i_{1}, j_{1}\right), \ldots,\left(i_{k}, j_{k}\right)\right)$, where, for $r=1,2, \ldots, k$, the numbers $i_{r}$ and $j_{r}$ are the indices of the two sheets that are chosen in the $r^{\text {th }}$ step. The resulting final sum will be some linear combination of $x_{1}, \ldots, x_{N}$, say, $c_{1} x_{1}+\cdots+c_{N} x_{N}$ with positive integers $c_{1}, \ldots, c_{N}$ that depend on $S$ only. Call the numbers $\left(c_{1}, \ldots, c_{N}\right)$ the characteristic vector of $S$.

Choose a sequence $S_{0}=\left(\left(i_{1}, j_{1}\right), \ldots,\left(i_{k}, j_{k}\right)\right)$ of steps that produces the minimal sum, starting from $x_{1}, \ldots, x_{N}$, and let $\left(c_{1}, \ldots, c_{N}\right)$ be the characteristic vector of $S$. We may assume that the sheets are indexed in such an order that $c_{1} \geqslant c_{2} \geqslant \cdots \geqslant c_{N}$. If the sheets (and the numbers) are permuted by a permutation $\pi$ of the indices $(1,2, \ldots, N)$ and then the same steps are performed, we can obtain the sum $\sum_{t=1}^{N} c_{t} x_{\pi(t)}$. By the rearrangement inequality, the smallest possible sum can be achieved when the numbers $\left(x_{1}, \ldots, x_{N}\right)$ are in non-decreasing order. So we can assume that also $x_{1} \leqslant x_{2} \leqslant \cdots \leqslant x_{N}$.

Let $\ell$ be the largest index with $c_{1}=\cdots=c_{\ell}$, and let the $r^{\text {th }}$ step be the first step for which $c_{i_{r}}=c_{1}$ or $c_{j_{r}}=c_{1}$. The role of $i_{r}$ and $j_{r}$ is symmetrical, so we can assume $c_{i_{r}}=c_{1}$ and thus $i_{r} \leqslant \ell$. We show that $c_{j_{r}}=c_{1}$ and $j_{r} \leqslant \ell$ hold, too.

Before the $r^{\text {th }}$ step, on the $i_{r}{ }^{\text {th }}$ sheet we had the number $x_{i_{r}}$. On the $j_{r}^{\text {th }}$ sheet there was a linear combination that contains the number $x_{j_{r}}$ with a positive integer coefficient, and possibly some other terms. In the $r^{\text {th }}$ step, the number $x_{i_{r}}$ joins that linear combination. From this point, each sheet contains a linear combination of $x_{1}, \ldots, x_{N}$, with the coefficient of $x_{j_{r}}$ being not smaller than the coefficient of $x_{i_{r}}$. This is preserved to the end of the procedure, so we have $c_{j_{r}} \geqslant c_{i_{r}}$. But $c_{i_{r}}=c_{1}$ is maximal among the coefficients, so we have $c_{j_{r}}=c_{i_{r}}=c_{1}$ and thus $j_{r} \leqslant \ell$.

Either from $c_{j_{r}}=c_{i_{r}}=c_{1}$ or from the arguments in the previous paragraph we can see that none of the $i_{r}{ }^{\text {th }}$ and the $j_{r}{ }^{\text {th }}$ sheets were used before step $r$. Therefore, the final linear combination of the numbers does not change if the step $\left(i_{r}, j_{r}\right)$ is performed first: the sequence of steps

$$
S_{1}=\left(\left(i_{r}, j_{r}\right),\left(i_{1}, j_{1}\right), \ldots,\left(i_{r-1}, j_{r-1}\right),\left(i_{r+1}, j_{r+1}\right), \ldots,\left(i_{N}, j_{N}\right)\right)
$$

also produces the same minimal sum at the end. Therefore, we can replace $S_{0}$ by $S_{1}$ and we may assume that $r=1$ and $c_{i_{1}}=c_{j_{1}}=c_{1}$.

As $i_{1} \neq j_{1}$, we can see that $\ell \geqslant 2$ and $c_{1}=c_{2}=c_{i_{1}}=c_{j_{1}}$. Let $\pi$ be such a permutation of the indices $(1,2, \ldots, N)$ that exchanges 1,2 with $i_{r}, j_{r}$ and does not change the remaining indices. Let

$$
S_{2}=\left(\left(\pi\left(i_{1}\right), \pi\left(j_{1}\right)\right), \ldots,\left(\pi\left(i_{N}\right), \pi\left(j_{N}\right)\right)\right) .
$$

Since $c_{\pi(i)}=c_{i}$ for all indices $i$, this sequence of steps produces the same, minimal sum. Moreover, in the first step we chose $x_{\pi\left(i_{1}\right)}=x_{1}$ and $x_{\pi\left(j_{1}\right)}=x_{2}$, the two smallest numbers.

Hence, it is possible to achieve the optimal sum if we follow the greedy strategy in the first step. By the induction hypothesis, following the greedy strategy in the remaining steps we achieve the optimal sum.

C3. Let $n \geqslant 2$ be an integer. Consider an $n \times n$ chessboard divided into $n^{2}$ unit squares. We call a configuration of $n$ rooks on this board happy if every row and every column contains exactly one rook. Find the greatest positive integer $k$ such that for every happy configuration of rooks, we can find a $k \times k$ square without a rook on any of its $k^{2}$ unit squares.
(Croatia)
Answer. $\lfloor\sqrt{n-1}\rfloor$.
Solution. Let $\ell$ be a positive integer. We will show that (i) if $n>\ell^{2}$ then each happy configuration contains an empty $\ell \times \ell$ square, but (ii) if $n \leqslant \ell^{2}$ then there exists a happy configuration not containing such a square. These two statements together yield the answer.
(i). Assume that $n>\ell^{2}$. Consider any happy configuration. There exists a row $R$ containing a rook in its leftmost square. Take $\ell$ consecutive rows with $R$ being one of them. Their union $U$ contains exactly $\ell$ rooks. Now remove the $n-\ell^{2} \geqslant 1$ leftmost columns from $U$ (thus at least one rook is also removed). The remaining part is an $\ell^{2} \times \ell$ rectangle, so it can be split into $\ell$ squares of size $\ell \times \ell$, and this part contains at most $\ell-1$ rooks. Thus one of these squares is empty.
(ii). Now we assume that $n \leqslant \ell^{2}$. Firstly, we will construct a happy configuration with no empty $\ell \times \ell$ square for the case $n=\ell^{2}$. After that we will modify it to work for smaller values of $n$.

Let us enumerate the rows from bottom to top as well as the columns from left to right by the numbers $0,1, \ldots, \ell^{2}-1$. Every square will be denoted, as usual, by the pair $(r, c)$ of its row and column numbers. Now we put the rooks on all squares of the form $(i \ell+j, j \ell+i)$ with $i, j=0,1, \ldots, \ell-1$ (the picture below represents this arrangement for $\ell=3$ ). Since each number from 0 to $\ell^{2}-1$ has a unique representation of the form $i \ell+j(0 \leqslant i, j \leqslant \ell-1)$, each row and each column contains exactly one rook.


Next, we show that each $\ell \times \ell$ square $A$ on the board contains a rook. Consider such a square $A$, and consider $\ell$ consecutive rows the union of which contains $A$. Let the lowest of these rows have number $p \ell+q$ with $0 \leqslant p, q \leqslant \ell-1$ (notice that $p \ell+q \leqslant \ell^{2}-\ell$ ). Then the rooks in this union are placed in the columns with numbers $q \ell+p,(q+1) \ell+p, \ldots,(\ell-1) \ell+p$, $p+1, \ell+(p+1), \ldots,(q-1) \ell+p+1$, or, putting these numbers in increasing order,

$$
p+1, \ell+(p+1), \ldots,(q-1) \ell+(p+1), q \ell+p,(q+1) \ell+p, \ldots,(\ell-1) \ell+p .
$$

One readily checks that the first number in this list is at most $\ell-1$ (if $p=\ell-1$, then $q=0$, and the first listed number is $q \ell+p=\ell-1$ ), the last one is at least $(\ell-1) \ell$, and the difference between any two consecutive numbers is at most $\ell$. Thus, one of the $\ell$ consecutive columns intersecting $A$ contains a number listed above, and the rook in this column is inside $A$, as required. The construction for $n=\ell^{2}$ is established.

It remains to construct a happy configuration of rooks not containing an empty $\ell \times \ell$ square for $n<\ell^{2}$. In order to achieve this, take the construction for an $\ell^{2} \times \ell^{2}$ square described above and remove the $\ell^{2}-n$ bottom rows together with the $\ell^{2}-n$ rightmost columns. We will have a rook arrangement with no empty $\ell \times \ell$ square, but several rows and columns may happen to be empty. Clearly, the number of empty rows is equal to the number of empty columns, so one can find a bijection between them, and put a rook on any crossing of an empty row and an empty column corresponding to each other.

Comment. Part (i) allows several different proofs. E.g., in the last paragraph of the solution, it suffices to deal only with the case $n=\ell^{2}+1$. Notice now that among the four corner squares, at least one is empty. So the rooks in its row and in its column are distinct. Now, deleting this row and column we obtain an $\ell^{2} \times \ell^{2}$ square with $\ell^{2}-1$ rooks in it. This square can be partitioned into $\ell^{2}$ squares of size $\ell \times \ell$, so one of them is empty.

C4. Construct a tetromino by attaching two $2 \times 1$ dominoes along their longer sides such that the midpoint of the longer side of one domino is a corner of the other domino. This construction yields two kinds of tetrominoes with opposite orientations. Let us call them Sand Z-tetrominoes, respectively.


Assume that a lattice polygon $P$ can be tiled with S-tetrominoes. Prove than no matter how we tile $P$ using only S- and Z-tetrominoes, we always use an even number of Z-tetrominoes.
(Hungary)
Solution 1. We may assume that polygon $P$ is the union of some squares of an infinite chessboard. Colour the squares of the chessboard with two colours as the figure below illustrates.


Observe that no matter how we tile $P$, any S-tetromino covers an even number of black squares, whereas any Z-tetromino covers an odd number of them. As $P$ can be tiled exclusively by S-tetrominoes, it contains an even number of black squares. But if some S-tetrominoes and some Z-tetrominoes cover an even number of black squares, then the number of Z-tetrominoes must be even.

Comment. An alternative approach makes use of the following two colourings, which are perhaps somewhat more natural:



Let $s_{1}$ and $s_{2}$ be the number of $S$-tetrominoes of the first and second type (as shown in the figure above) respectively that are used in a tiling of $P$. Likewise, let $z_{1}$ and $z_{2}$ be the number of $Z$-tetrominoes of the first and second type respectively. The first colouring shows that $s_{1}+z_{2}$ is invariant modulo 2 , the second colouring shows that $s_{1}+z_{1}$ is invariant modulo 2 . Adding these two conditions, we find that $z_{1}+z_{2}$ is invariant modulo 2 , which is what we have to prove. Indeed, the sum of the two colourings (regarding white as 0 and black as 1 and adding modulo 2 ) is the colouring shown in the solution.

Solution 2. Let us assign coordinates to the squares of the infinite chessboard in such a way that the squares of $P$ have nonnegative coordinates only, and that the first coordinate increases as one moves to the right, while the second coordinate increases as one moves upwards. Write the integer $3^{i} \cdot(-3)^{j}$ into the square with coordinates $(i, j)$, as in the following figure:

| $\vdots$ |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 81 | $\vdots$ |  |  |  |  |
| -27 | -81 | $\vdots$ |  |  |  |
| 9 | 27 | 81 | $\cdots$ |  |  |
| -3 | -9 | -27 | -81 | $\cdots$ |  |
| 1 | 3 | 9 | 27 | 81 | $\cdots$ |

The sum of the numbers written into four squares that can be covered by an $S$-tetromino is either of the form

$$
3^{i} \cdot(-3)^{j} \cdot\left(1+3+3 \cdot(-3)+3^{2} \cdot(-3)\right)=-32 \cdot 3^{i} \cdot(-3)^{j}
$$

(for the first type of $S$-tetrominoes), or of the form

$$
3^{i} \cdot(-3)^{j} \cdot\left(3+3 \cdot(-3)+(-3)+(-3)^{2}\right)=0
$$

and thus divisible by 32 . For this reason, the sum of the numbers written into the squares of $P$, and thus also the sum of the numbers covered by $Z$-tetrominoes in the second covering, is likewise divisible by 32 . Now the sum of the entries of a $Z$-tetromino is either of the form

$$
3^{i} \cdot(-3)^{j} \cdot\left(3+3^{2}+(-3)+3 \cdot(-3)\right)=0
$$

(for the first type of $Z$-tetrominoes), or of the form

$$
3^{i} \cdot(-3)^{j} \cdot\left(1+(-3)+3 \cdot(-3)+3 \cdot(-3)^{2}\right)=16 \cdot 3^{i} \cdot(-3)^{j}
$$

i.e., 16 times an odd number. Thus in order to obtain a total that is divisible by 32 , an even number of the latter kind of $Z$-tetrominoes needs to be used. Rotating everything by $90^{\circ}$, we find that the number of $Z$-tetrominoes of the first kind is even as well. So we have even proven slightly more than necessary.

Comment 1. In the second solution, 3 and -3 can be replaced by other combinations as well. For example, for any positive integer $a \equiv 3(\bmod 4)$, we can write $a^{i} \cdot(-a)^{j}$ into the square with coordinates $(i, j)$ and apply the same argument.

Comment 2. As the second solution shows, we even have the stronger result that the parity of the number of each of the four types of tetrominoes in a tiling of $P$ by S- and Z-tetrominoes is an invariant of $P$. This also remains true if there is no tiling of $P$ that uses only S -tetrominoes.

C5. Consider $n \geqslant 3$ lines in the plane such that no two lines are parallel and no three have a common point. These lines divide the plane into polygonal regions; let $\mathcal{F}$ be the set of regions having finite area. Prove that it is possible to colour $\lceil\sqrt{n / 2}\rceil$ of the lines blue in such a way that no region in $\mathcal{F}$ has a completely blue boundary. (For a real number $x,\lceil x\rceil$ denotes the least integer which is not smaller than $x$.)
(Austria)
Solution. Let $L$ be the given set of lines. Choose a maximal (by inclusion) subset $B \subseteq L$ such that when we colour the lines of $B$ blue, no region in $\mathcal{F}$ has a completely blue boundary. Let $|B|=k$. We claim that $k \geqslant\lceil\sqrt{n / 2}\rceil$.

Let us colour all the lines of $L \backslash B$ red. Call a point blue if it is the intersection of two blue lines. Then there are $\binom{k}{2}$ blue points.

Now consider any red line $\ell$. By the maximality of $B$, there exists at least one region $A \in \mathcal{F}$ whose only red side lies on $\ell$. Since $A$ has at least three sides, it must have at least one blue vertex. Let us take one such vertex and associate it to $\ell$.

Since each blue point belongs to four regions (some of which may be unbounded), it is associated to at most four red lines. Thus the total number of red lines is at most $4\binom{k}{2}$. On the other hand, this number is $n-k$, so

$$
n-k \leqslant 2 k(k-1), \quad \text { thus } \quad n \leqslant 2 k^{2}-k \leqslant 2 k^{2},
$$

and finally $k \geqslant\lceil\sqrt{n / 2}\rceil$, which gives the desired result.

Comment 1. The constant factor in the estimate can be improved in different ways; we sketch two of them below. On the other hand, the Problem Selection Committee is not aware of any results showing that it is sometimes impossible to colour $k$ lines satisfying the desired condition for $k \gg \sqrt{n}$. In this situation we find it more suitable to keep the original formulation of the problem.

1. Firstly, we show that in the proof above one has in fact $k=|B| \geqslant\lceil\sqrt{2 n / 3}\rceil$.

Let us make weighted associations as follows. Let a region $A$ whose only red side lies on $\ell$ have $k$ vertices, so that $k-2$ of them are blue. We associate each of these blue vertices to $\ell$, and put the weight $\frac{1}{k-2}$ on each such association. So the sum of the weights of all the associations is exactly $n-k$.

Now, one may check that among the four regions adjacent to a blue vertex $v$, at most two are triangles. This means that the sum of the weights of all associations involving $v$ is at most $1+1+\frac{1}{2}+\frac{1}{2}=3$. This leads to the estimate

$$
n-k \leqslant 3\binom{k}{2},
$$

or

$$
2 n \leqslant 3 k^{2}-k<3 k^{2}
$$

which yields $k \geqslant\lceil\sqrt{2 n / 3}\rceil$.
2. Next, we even show that $k=|B| \geqslant\lceil\sqrt{n}\rceil$. For this, we specify the process of associating points to red lines in one more different way.

Call a point red if it lies on a red line as well as on a blue line. Consider any red line $\ell$, and take an arbitrary region $A \in \mathcal{F}$ whose only red side lies on $\ell$. Let $r^{\prime}, r, b_{1}, \ldots, b_{k}$ be its vertices in clockwise order with $r^{\prime}, r \in \ell$; then the points $r^{\prime}, r$ are red, while all the points $b_{1}, \ldots, b_{k}$ are blue. Let us associate to $\ell$ the red point $r$ and the blue point $b_{1}$. One may notice that to each pair of a red point $r$ and a blue point $b$, at most one red line can be associated, since there is at most one region $A$ having $r$ and $b$ as two clockwise consecutive vertices.

We claim now that at most two red lines are associated to each blue point $b$; this leads to the desired bound

$$
n-k \leqslant 2\binom{k}{2} \quad \Longleftrightarrow \quad n \leqslant k^{2}
$$

Assume, to the contrary, that three red lines $\ell_{1}, \ell_{2}$, and $\ell_{3}$ are associated to the same blue point $b$. Let $r_{1}, r_{2}$, and $r_{3}$ respectively be the red points associated to these lines; all these points are distinct. The point $b$ defines four blue rays, and each point $r_{i}$ is the red point closest to $b$ on one of these rays. So we may assume that the points $r_{2}$ and $r_{3}$ lie on one blue line passing through $b$, while $r_{1}$ lies on the other one.


Now consider the region $A$ used to associate $r_{1}$ and $b$ with $\ell_{1}$. Three of its clockwise consecutive vertices are $r_{1}, b$, and either $r_{2}$ or $r_{3}$ (say, $r_{2}$ ). Since $A$ has only one red side, it can only be the triangle $r_{1} b r_{2}$; but then both $\ell_{1}$ and $\ell_{2}$ pass through $r_{2}$, as well as some blue line. This is impossible by the problem assumptions.

Comment 2. The condition that the lines be non-parallel is essentially not used in the solution, nor in the previous comment; thus it may be omitted.

C6. We are given an infinite deck of cards, each with a real number on it. For every real number $x$, there is exactly one card in the deck that has $x$ written on it. Now two players draw disjoint sets $A$ and $B$ of 100 cards each from this deck. We would like to define a rule that declares one of them a winner. This rule should satisfy the following conditions:

1. The winner only depends on the relative order of the 200 cards: if the cards are laid down in increasing order face down and we are told which card belongs to which player, but not what numbers are written on them, we can still decide the winner.
2. If we write the elements of both sets in increasing order as $A=\left\{a_{1}, a_{2}, \ldots, a_{100}\right\}$ and $B=\left\{b_{1}, b_{2}, \ldots, b_{100}\right\}$, and $a_{i}>b_{i}$ for all $i$, then $A$ beats $B$.
3. If three players draw three disjoint sets $A, B, C$ from the deck, $A$ beats $B$ and $B$ beats $C$, then $A$ also beats $C$.

How many ways are there to define such a rule? Here, we consider two rules as different if there exist two sets $A$ and $B$ such that $A$ beats $B$ according to one rule, but $B$ beats $A$ according to the other.
(Russia)
Answer. 100.
Solution 1. We prove a more general statement for sets of cardinality $n$ (the problem being the special case $n=100$, then the answer is $n$ ). In the following, we write $A>B$ or $B<A$ for " $A$ beats $B$ ".

Part I. Let us first define $n$ different rules that satisfy the conditions. To this end, fix an index $k \in\{1,2, \ldots, n\}$. We write both $A$ and $B$ in increasing order as $A=\left\{a_{1}, a_{2}, \ldots, a_{n}\right\}$ and $B=\left\{b_{1}, b_{2}, \ldots, b_{n}\right\}$ and say that $A$ beats $B$ if and only if $a_{k}>b_{k}$. This rule clearly satisfies all three conditions, and the rules corresponding to different $k$ are all different. Thus there are at least $n$ different rules.

Part II. Now we have to prove that there is no other way to define such a rule. Suppose that our rule satisfies the conditions, and let $k \in\{1,2, \ldots, n\}$ be minimal with the property that

$$
A_{k}=\{1,2, \ldots, k, n+k+1, n+k+2, \ldots, 2 n\}<B_{k}=\{k+1, k+2, \ldots, n+k\} .
$$

Clearly, such a $k$ exists, since this holds for $k=n$ by assumption. Now consider two disjoint sets $X=\left\{x_{1}, x_{2}, \ldots, x_{n}\right\}$ and $Y=\left\{y_{1}, y_{2}, \ldots, y_{n}\right\}$, both in increasing order (i.e., $x_{1}<x_{2}<\cdots<x_{n}$ and $y_{1}<y_{2}<\cdots<y_{n}$ ). We claim that $X<Y$ if (and only if - this follows automatically) $x_{k}<y_{k}$.

To prove this statement, pick arbitrary real numbers $u_{i}, v_{i}, w_{i} \notin X \cup Y$ such that

$$
u_{1}<u_{2}<\cdots<u_{k-1}<\min \left(x_{1}, y_{1}\right), \quad \max \left(x_{n}, y_{n}\right)<v_{k+1}<v_{k+2}<\cdots<v_{n}
$$

and

$$
x_{k}<v_{1}<v_{2}<\cdots<v_{k}<w_{1}<w_{2}<\cdots<w_{n}<u_{k}<u_{k+1}<\cdots<u_{n}<y_{k}
$$

and set

$$
U=\left\{u_{1}, u_{2}, \ldots, u_{n}\right\}, V=\left\{v_{1}, v_{2}, \ldots, v_{n}\right\}, W=\left\{w_{1}, w_{2}, \ldots, w_{n}\right\} .
$$

Then

- $u_{i}<y_{i}$ and $x_{i}<v_{i}$ for all $i$, so $U<Y$ and $X<V$ by the second condition.
- The elements of $U \cup W$ are ordered in the same way as those of $A_{k-1} \cup B_{k-1}$, and since $A_{k-1}>B_{k-1}$ by our choice of $k$, we also have $U>W$ (if $k=1$, this is trivial).
- The elements of $V \cup W$ are ordered in the same way as those of $A_{k} \cup B_{k}$, and since $A_{k}<B_{k}$ by our choice of $k$, we also have $V<W$.

It follows that

$$
X<V<W<U<Y
$$

so $X<Y$ by the third condition, which is what we wanted to prove.
Solution 2. Another possible approach to Part II of this problem is induction on $n$. For $n=1$, there is trivially only one rule in view of the second condition.

In the following, we assume that our claim (namely, that there are no possible rules other than those given in Part I) holds for $n-1$ in place of $n$. We start with the following observation: Claim. At least one of the two relations

$$
(\{2\} \cup\{2 i-1 \mid 2 \leqslant i \leqslant n\})<(\{1\} \cup\{2 i \mid 2 \leqslant i \leqslant n\})
$$

and

$$
(\{2 i-1 \mid 1 \leqslant i \leqslant n-1\} \cup\{2 n\})<(\{2 i \mid 1 \leqslant i \leqslant n-1\} \cup\{2 n-1\})
$$

holds.
Proof. Suppose that the first relation does not hold. Since our rule may only depend on the relative order, we must also have

$$
(\{2\} \cup\{3 i-2 \mid 2 \leqslant i \leqslant n-1\} \cup\{3 n-2\})>(\{1\} \cup\{3 i-1 \mid 2 \leqslant i \leqslant n-1\} \cup\{3 n\}) .
$$

Likewise, if the second relation does not hold, then we must also have

$$
(\{1\} \cup\{3 i-1 \mid 2 \leqslant i \leqslant n-1\} \cup\{3 n\})>(\{3\} \cup\{3 i \mid 2 \leqslant i \leqslant n-1\} \cup\{3 n-1\}) .
$$

Now condition 3 implies that

$$
(\{2\} \cup\{3 i-2 \mid 2 \leqslant i \leqslant n-1\} \cup\{3 n-2\})>(\{3\} \cup\{3 i \mid 2 \leqslant i \leqslant n-1\} \cup\{3 n-1\}),
$$

which contradicts the second condition.
Now we distinguish two cases, depending on which of the two relations actually holds:
First case: $(\{2\} \cup\{2 i-1 \mid 2 \leqslant i \leqslant n\})<(\{1\} \cup\{2 i \mid 2 \leqslant i \leqslant n\})$.
Let $A=\left\{a_{1}, a_{2}, \ldots, a_{n}\right\}$ and $B=\left\{b_{1}, b_{2}, \ldots, b_{n}\right\}$ be two disjoint sets, both in increasing order. We claim that the winner can be decided only from the values of $a_{2}, \ldots, a_{n}$ and $b_{2}, \ldots, b_{n}$, while $a_{1}$ and $b_{1}$ are actually irrelevant. Suppose that this was not the case, and assume without loss of generality that $a_{2}<b_{2}$. Then the relative order of $a_{1}, a_{2}, \ldots, a_{n}, b_{2}, \ldots, b_{n}$ is fixed, and the position of $b_{1}$ has to decide the winner. Suppose that for some value $b_{1}=x, B$ wins, while for some other value $b_{1}=y, A$ wins.

Write $B_{x}=\left\{x, b_{2}, \ldots, b_{n}\right\}$ and $B_{y}=\left\{y, b_{2}, \ldots, b_{n}\right\}$, and let $\varepsilon>0$ be smaller than half the distance between any two of the numbers in $B_{x} \cup B_{y} \cup A$. For any set $M$, let $M \pm \varepsilon$ be the set obtained by adding/subtracting $\varepsilon$ to all elements of $M$. By our choice of $\varepsilon$, the relative order of the elements of $\left(B_{y}+\varepsilon\right) \cup A$ is still the same as for $B_{y} \cup A$, while the relative order of the elements of $\left(B_{x}-\varepsilon\right) \cup A$ is still the same as for $B_{x} \cup A$. Thus $A<B_{x}-\varepsilon$, but $A>B_{y}+\varepsilon$. Moreover, if $y>x$, then $B_{x}-\varepsilon<B_{y}+\varepsilon$ by condition 2 , while otherwise the relative order of
the elements in $\left(B_{x}-\varepsilon\right) \cup\left(B_{y}+\varepsilon\right)$ is the same as for the two sets $\{2\} \cup\{2 i-1 \mid 2 \leqslant i \leqslant n\}$ and $\{1\} \cup\{2 i \mid 2 \leqslant i \leqslant n\}$, so that $B_{x}-\varepsilon<B_{y}+\varepsilon$. In either case, we obtain

$$
A<B_{x}-\varepsilon<B_{y}+\varepsilon<A,
$$

which contradicts condition 3 .
So we know now that the winner does not depend on $a_{1}, b_{1}$. Therefore, we can define a new rule $<^{*}$ on sets of cardinality $n-1$ by saying that $A<^{*} B$ if and only if $A \cup\{a\}<B \cup\{b\}$ for some $a, b$ (or equivalently, all $a, b$ ) such that $a<\min A, b<\min B$ and $A \cup\{a\}$ and $B \cup\{b\}$ are disjoint. The rule $<^{*}$ satisfies all conditions again, so by the induction hypothesis, there exists an index $i$ such that $A<^{*} B$ if and only if the $i^{\text {th }}$ smallest element of $A$ is less than the $i^{\text {th }}$ smallest element of $B$. This implies that $C<D$ if and only if the $(i+1)^{\text {th }}$ smallest element of $C$ is less than the $(i+1)^{\text {th }}$ smallest element of $D$, which completes our induction.

Second case: $(\{2 i-1 \mid 1 \leqslant i \leqslant n-1\} \cup\{2 n\})<(\{2 i \mid 1 \leqslant i \leqslant n-1\} \cup\{2 n-1\})$.
Set $-A=\{-a \mid a \in A\}$ for any $A \subseteq \mathbb{R}$. For any two disjoint sets $A, B \subseteq \mathbb{R}$ of cardinality $n$, we write $A \prec^{\circ} B$ to mean $(-B)<(-A)$. It is easy to see that $<^{\circ}$ defines a rule to determine a winner that satisfies the three conditions of our problem as well as the relation of the first case. So it follows in the same way as in the first case that for some $i, A<^{\circ} B$ if and only if the $i^{\text {th }}$ smallest element of $A$ is less than the $i^{\text {th }}$ smallest element of $B$, which is equivalent to the condition that the $i^{\text {th }}$ largest element of $-A$ is greater than the $i^{\text {th }}$ largest element of $-B$. This proves that the original rule $<$ also has the desired form.

Comment. The problem asks for all possible partial orders on the set of $n$-element subsets of $\mathbb{R}$ such that any two disjoint sets are comparable, the order relation only depends on the relative order of the elements, and $\left\{a_{1}, a_{2}, \ldots, a_{n}\right\}<\left\{b_{1}, b_{2}, \ldots, b_{n}\right\}$ whenever $a_{i}<b_{i}$ for all $i$.

As the proposer points out, one may also ask for all total orders on all $n$-element subsets of $\mathbb{R}$ (dropping the condition of disjointness and requiring that $\left\{a_{1}, a_{2}, \ldots, a_{n}\right\} \leq\left\{b_{1}, b_{2}, \ldots, b_{n}\right\}$ whenever $a_{i} \leqslant b_{i}$ for all $i$ ). It turns out that the number of possibilities in this case is $n!$, and all possible total orders are obtained in the following way. Fix a permutation $\sigma \in S_{n}$. Let $A=\left\{a_{1}, a_{2}, \ldots, a_{n}\right\}$ and $B=\left\{b_{1}, b_{2}, \ldots, b_{n}\right\}$ be two subsets of $\mathbb{R}$ with $a_{1}<a_{2}<\cdots<a_{n}$ and $b_{1}<b_{2}<\cdots<b_{n}$. Then we say that $A>_{\sigma} B$ if and only if $\left(a_{\sigma(1)}, \ldots, a_{\sigma(n)}\right)$ is lexicographically greater than $\left(b_{\sigma(1)}, \ldots, b_{\sigma(n)}\right)$.

It seems, however, that this formulation adds rather more technicalities to the problem than additional ideas.

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C7. Let $M$ be a set of $n \geqslant 4$ points in the plane, no three of which are collinear. Initially these points are connected with $n$ segments so that each point in $M$ is the endpoint of exactly two segments. Then, at each step, one may choose two segments $A B$ and $C D$ sharing a common interior point and replace them by the segments $A C$ and $B D$ if none of them is present at this moment. Prove that it is impossible to perform $n^{3} / 4$ or more such moves.
(Russia)
Solution. A line is said to be red if it contains two points of $M$. As no three points of $M$ are collinear, each red line determines a unique pair of points of $M$. Moreover, there are precisely $\binom{n}{2}<\frac{n^{2}}{2}$ red lines. By the value of a segment we mean the number of red lines intersecting it in its interior, and the value of a set of segments is defined to be the sum of the values of its elements. We will prove that $(i)$ the value of the initial set of segments is smaller than $n^{3} / 2$ and that (ii) each step decreases the value of the set of segments present by at least 2 . Since such a value can never be negative, these two assertions imply the statement of the problem.

To show ( $i$ ) we just need to observe that each segment has a value that is smaller than $n^{2} / 2$. Thus the combined value of the $n$ initial segments is indeed below $n \cdot n^{2} / 2=n^{3} / 2$.

It remains to establish (ii). Suppose that at some moment we have two segments $A B$ and $C D$ sharing an interior point $S$, and that at the next moment we have the two segments $A C$ and $B D$ instead. Let $X_{A B}$ denote the set of red lines intersecting the segment $A B$ in its interior and let the sets $X_{A C}, X_{B D}$, and $X_{C D}$ be defined similarly. We are to prove that $\left|X_{A C}\right|+\left|X_{B D}\right|+2 \leqslant\left|X_{A B}\right|+\left|X_{C D}\right|$.

As a first step in this direction, we claim that

$$
\begin{equation*}
\left|X_{A C} \cup X_{B D}\right|+2 \leqslant\left|X_{A B} \cup X_{C D}\right| . \tag{1}
\end{equation*}
$$

Indeed, if $g$ is a red line intersecting, e.g. the segment $A C$ in its interior, then it has to intersect the triangle $A C S$ once again, either in the interior of its side $A S$, or in the interior of its side $C S$, or at $S$, meaning that it belongs to $X_{A B}$ or to $X_{C D}$ (see Figure 1). Moreover, the red lines $A B$ and $C D$ contribute to $X_{A B} \cup X_{C D}$ but not to $X_{A C} \cup X_{B D}$. Thereby (1) is proved.


Similarly but more easily one obtains

$$
\begin{equation*}
\left|X_{A C} \cap X_{B D}\right| \leqslant\left|X_{A B} \cap X_{C D}\right| \tag{2}
\end{equation*}
$$

Indeed, a red line $h$ appearing in $X_{A C} \cap X_{B D}$ belongs, for similar reasons as above, also to $X_{A B} \cap X_{C D}$. To make the argument precise, one may just distinguish the cases $S \in h$ (see Figure 2) and $S \notin h$ (see Figure 3). Thereby (2) is proved.

Adding (1) and (2) we obtain the desired conclusion, thus completing the solution of this problem.

Comment 1. There is a problem belonging to the folklore, in the solution of which one may use the same kind of operation:

Given $n$ red and $n$ green points in the plane, prove that one may draw $n$ nonintersecting segments each of which connects a red point with a green point.

A standard approach to this problem consists in taking $n$ arbitrary segments connecting the red points with the green points, and to perform the same operation as in the above proposal whenever an intersection occurs. Now each time one performs such a step, the total length of the segments that are present decreases due to the triangle inequality. So, as there are only finitely many possibilities for the set of segments present, the process must end at some stage.

In the above proposal, however, considering the sum of the Euclidean lengths of the segment that are present does not seem to help much, for even though it shows that the process must necessarily terminate after some finite number of steps, it does not seem to easily yield any upper bound on the number of these steps that grows polynomially with $n$.

One may regard the concept of the value of a segment introduced in the above solution as an appropriately discretised version of Euclidean length suitable for obtaining such a bound.

The Problem Selection Committee still believes the problem to be sufficiently original for the competition.

Comment 2. There are some other essentially equivalent ways of presenting the same solution. E.g., put $M=\left\{A_{1}, A_{2}, \ldots, A_{n}\right\}$, denote the set of segments present at any moment by $\left\{e_{1}, e_{2}, \ldots, e_{n}\right\}$, and called a triple $(i, j, k)$ of indices with $i \neq j$ intersecting, if the line $A_{i} A_{j}$ intersects the segment $e_{k}$. It may then be shown that the number $S$ of intersecting triples satisfies $0 \leqslant S<n^{3}$ at the beginning and decreases by at least 4 in each step.

Comment 3. It is not difficult to construct an example where $c n^{2}$ moves are possible (for some absolute constant $c>0$ ). It would be interesting to say more about the gap between $c n^{2}$ and $c n^{3}$.

C8. A card deck consists of 1024 cards. On each card, a set of distinct decimal digits is written in such a way that no two of these sets coincide (thus, one of the cards is empty). Two players alternately take cards from the deck, one card per turn. After the deck is empty, each player checks if he can throw out one of his cards so that each of the ten digits occurs on an even number of his remaining cards. If one player can do this but the other one cannot, the one who can is the winner; otherwise a draw is declared.

Determine all possible first moves of the first player after which he has a winning strategy.
(Russia)
Answer. All the moves except for taking the empty card.

Solution. Let us identify each card with the set of digits written on it. For any collection of cards $C_{1}, C_{2}, \ldots, C_{k}$ denote by their sum the set $C_{1} \triangle C_{2} \triangle \cdots \Delta C_{k}$ consisting of all elements belonging to an odd number of the $C_{i}$ 's. Denote the first and the second player by $\mathcal{F}$ and $\mathcal{S}$, respectively.

Since each digit is written on exactly 512 cards, the sum of all the cards is $\varnothing$. Therefore, at the end of the game the sum of all the cards of $\mathcal{F}$ will be the same as that of $\mathcal{S}$; denote this sum by $C$. Then the player who took $C$ can throw it out and get the desired situation, while the other one cannot. Thus, the player getting card $C$ wins, and no draw is possible.

Now, given a nonempty card $B$, one can easily see that all the cards can be split into 512 pairs of the form $(X, X \triangle B)$ because $(X \triangle B) \triangle B=X$. The following lemma shows a property of such a partition that is important for the solution.
Lemma. Let $B \neq \varnothing$ be some card. Let us choose 512 cards so that exactly one card is chosen from every pair $(X, X \triangle B)$. Then the sum of all chosen cards is either $\varnothing$ or $B$.
Proof. Let $b$ be some element of $B$. Enumerate the pairs; let $X_{i}$ be the card not containing $b$ in the $i^{\text {th }}$ pair, and let $Y_{i}$ be the other card in this pair. Then the sets $X_{i}$ are exactly all the sets not containing $b$, therefore each digit $a \neq b$ is written on exactly 256 of these cards, so $X_{1} \triangle X_{2} \triangle \cdots \Delta X_{512}=\varnothing$. Now, if we replace some summands in this sum by the other elements from their pairs, we will simply add $B$ several times to this sum, thus the sum will either remain unchanged or change by $B$, as required.

Now we consider two cases.
Case 1. Assume that $\mathcal{F}$ takes the card $\varnothing$ on his first move. In this case, we present a winning strategy for $\mathcal{S}$.

Let $\mathcal{S}$ take an arbitrary card $A$. Assume that $\mathcal{F}$ takes card $B$ after that; then $\mathcal{S}$ takes $A \triangle B$. Split all 1024 cards into 512 pairs of the form $(X, X \triangle B)$; we call two cards in one pair partners. Then the four cards taken so far form two pairs $(\varnothing, B)$ and $(A, A \triangle B)$ belonging to $\mathcal{F}$ and $\mathcal{S}$, respectively. On each of the subsequent moves, when $\mathcal{F}$ takes some card, $\mathcal{S}$ should take the partner of this card in response.

Consider the situation at the end of the game. Let us for a moment replace card $A$ belonging to $\mathcal{S}$ by $\varnothing$. Then he would have one card from each pair; by our lemma, the sum of all these cards would be either $\varnothing$ or $B$. Now, replacing $\varnothing$ back by $A$ we get that the actual sum of the cards of $\mathcal{S}$ is either $A$ or $A \triangle B$, and he has both these cards. Thus $\mathcal{S}$ wins.

Case 2. Now assume that $\mathcal{F}$ takes some card $A \neq \varnothing$ on his first move. Let us present $a$ winning strategy for $\mathcal{F}$ in this case.

Assume that $\mathcal{S}$ takes some card $B \neq \varnothing$ on his first move; then $\mathcal{F}$ takes $A \triangle B$. Again, let us split all the cards into pairs of the form $(X, X \triangle B)$; then the cards which have not been taken yet form several complete pairs and one extra element (card $\varnothing$ has not been taken while its partner $B$ has). Now, on each of the subsequent moves, if $\mathcal{S}$ takes some element from a
complete pair, then $\mathcal{F}$ takes its partner. If $\mathcal{S}$ takes the extra element, then $\mathcal{F}$ takes an arbitrary card $Y$, and the partner of $Y$ becomes the new extra element.

Thus, on his last move $\mathcal{S}$ is forced to take the extra element. After that player $\mathcal{F}$ has cards $A$ and $A \triangle B$, player $\mathcal{S}$ has cards $B$ and $\varnothing$, and $\mathcal{F}$ has exactly one element from every other pair. Thus the situation is the same as in the previous case with roles reversed, and $\mathcal{F}$ wins.

Finally, if $\mathcal{S}$ takes $\varnothing$ on his first move then $\mathcal{F}$ denotes any card which has not been taken yet by $B$ and takes $A \triangle B$. After that, the same strategy as above is applicable.

Comment 1. If one wants to avoid the unusual question about the first move, one may change the formulation as follows. (The difficulty of the problem would decrease somewhat.)

A card deck consists of 1023 cards; on each card, a nonempty set of distinct decimal digits is written in such a way that no two of these sets coincide. Two players alternately take cards from the deck, one card per turn. When the deck is empty, each player checks if he can throw out one of his cards so that for each of the ten digits, he still holds an even number of cards with this digit. If one player can do this but the other one cannot, the one who can is the winner; otherwise a draw is declared.

Determine which of the players (if any) has a winning strategy.
The winner in this version is the first player. The analysis of the game from the first two paragraphs of the previous solution applies to this version as well, except for the case $C=\varnothing$ in which the result is a draw. Then the strategy for $\mathcal{S}$ in Case 1 works for $\mathcal{F}$ in this version: the sum of all his cards at the end is either $A$ or $A \triangle B$, thus nonempty in both cases.

Comment 2. Notice that all the cards form a vector space over $\mathbb{F}_{2}$, with $\triangle$ the operation of addition. Due to the automorphisms of this space, all possibilities for $\mathcal{F}$ 's first move except $\varnothing$ are equivalent. The same holds for the response by $\mathcal{S}$ if $\mathcal{F}$ takes the card $\varnothing$ on his first move.

Comment 3. It is not that hard to show that in the initial game, $\mathcal{F}$ has a winning move, by the idea of "strategy stealing".

Namely, assume that $\mathcal{S}$ has a winning strategy. Let us take two card decks and start two games, in which $\mathcal{S}$ will act by his strategy. In the first game, $\mathcal{F}$ takes an arbitrary card $A_{1}$; assume that $\mathcal{S}$ takes some $B_{1}$ in response. Then $\mathcal{F}$ takes the card $B_{1}$ at the second game; let the response by $\mathcal{S}$ be $A_{2}$. Then $\mathcal{F}$ takes $A_{2}$ in the first game and gets a response $B_{2}$, and so on.

This process stops at some moment when in the second game $\mathcal{S}$ takes $A_{i}=A_{1}$. At this moment the players hold the same sets of cards in both games, but with roles reversed. Now, if some cards remain in the decks, $\mathcal{F}$ takes an arbitrary card from the first deck starting a similar cycle.

At the end of the game, player $\mathcal{F}$ 's cards in the first game are exactly player $\mathcal{S}$ 's cards in the second game, and vice versa. Thus in one of the games $\mathcal{F}$ will win, which is impossible by our assumption.

One may notice that the strategy in Case 2 is constructed exactly in this way from the strategy in Case 1 . This is possible since every response by $\mathcal{S}$ wins if $\mathcal{F}$ takes the card $\varnothing$ on his first move.

C9. There are $n$ circles drawn on a piece of paper in such a way that any two circles intersect in two points, and no three circles pass through the same point. Turbo the snail slides along the circles in the following fashion. Initially he moves on one of the circles in clockwise direction. Turbo always keeps sliding along the current circle until he reaches an intersection with another circle. Then he continues his journey on this new circle and also changes the direction of moving, i.e. from clockwise to anticlockwise or vice versa.

Suppose that Turbo's path entirely covers all circles. Prove that $n$ must be odd.
(India)
Solution 1. Replace every cross (i.e. intersection of two circles) by two small circle arcs that indicate the direction in which the snail should leave the cross (see Figure 1.1). Notice that the placement of the small arcs does not depend on the direction of moving on the curves; no matter which direction the snail is moving on the circle arcs, he will follow the same curves (see Figure 1.2). In this way we have a set of curves, that are the possible paths of the snail. Call these curves snail orbits or just orbits. Every snail orbit is a simple closed curve that has no intersection with any other orbit.


Figure 1.1


Figure 1.2

We prove the following, more general statement.
(*) In any configuration of $n$ circles such that no two of them are tangent, the number of snail orbits has the same parity as the number $n$. (Note that it is not assumed that all circle pairs intersect.)

This immediately solves the problem.
Let us introduce the following operation that will be called fipping a cross. At a cross, remove the two small arcs of the orbits, and replace them by the other two arcs. Hence, when the snail arrives at a flipped cross, he will continue on the other circle as before, but he will preserve the orientation in which he goes along the circle arcs (see Figure 2).

b


Figure 2
Consider what happens to the number of orbits when a cross is flipped. Denote by $a, b, c$, and $d$ the four arcs that meet at the cross such that $a$ and $b$ belong to the same circle. Before the flipping $a$ and $b$ were connected to $c$ and $d$, respectively, and after the flipping $a$ and $b$ are connected to $d$ and $c$, respectively.

The orbits passing through the cross are closed curves, so each of the arcs $a, b, c$, and $d$ is connected to another one by orbits outside the cross. We distinguish three cases.

Case 1: $a$ is connected to $b$ and $c$ is connected to $d$ by the orbits outside the cross (see Figure 3.1).

We show that this case is impossible. Remove the two small arcs at the cross, connect $a$ to $b$, and connect $c$ to $d$ at the cross. Let $\gamma$ be the new closed curve containing $a$ and $b$, and let $\delta$ be the new curve that connects $c$ and $d$. These two curves intersect at the cross. So one of $c$ and $d$ is inside $\gamma$ and the other one is outside $\gamma$. Then the two closed curves have to meet at least one more time, but this is a contradiction, since no orbit can intersect itself.


Figure 3.1


Figure 3.2


Figure 3.3

Case 2: $a$ is connected to $c$ and $b$ is connected to $d$ (see Figure 3.2).
Before the flipping $a$ and $c$ belong to one orbit and $b$ and $d$ belong to another orbit. Flipping the cross merges the two orbits into a single orbit. Hence, the number of orbits decreases by 1.

Case 3: $a$ is connected to $d$ and $b$ is connected to $c$ (see Figure 3.3).
Before the flipping the arcs $a, b, c$, and $d$ belong to a single orbit. Flipping the cross splits that orbit in two. The number of orbits increases by 1 .

As can be seen, every flipping decreases or increases the number of orbits by one, thus changes its parity.

Now flip every cross, one by one. Since every pair of circles has 0 or 2 intersections, the number of crosses is even. Therefore, when all crosses have been flipped, the original parity of the number of orbits is restored. So it is sufficient to prove (*) for the new configuration, where all crosses are flipped. Of course also in this new configuration the (modified) orbits are simple closed curves not intersecting each other.

Orient the orbits in such a way that the snail always moves anticlockwise along the circle arcs. Figure 4 shows the same circles as in Figure 1 after flipping all crosses and adding orientation. (Note that this orientation may be different from the orientation of the orbit as a planar curve; the orientation of every orbit may be negative as well as positive, like the middle orbit in Figure 4.) If the snail moves around an orbit, the total angle change in his moving direction, the total curvature, is either $+2 \pi$ or $-2 \pi$, depending on the orientation of the orbit. Let $P$ and $N$ be the number of orbits with positive and negative orientation, respectively. Then the total curvature of all orbits is $(P-N) \cdot 2 \pi$.


Figure 4


Figure 5

Double-count the total curvature of all orbits. Along every circle the total curvature is $2 \pi$. At every cross, the two turnings make two changes with some angles having the same absolute value but opposite signs, as depicted in Figure 5. So the changes in the direction at the crosses cancel out. Hence, the total curvature is $n \cdot 2 \pi$.

Now we have $(P-N) \cdot 2 \pi=n \cdot 2 \pi$, so $P-N=n$. The number of (modified) orbits is $P+N$, that has a same parity as $P-N=n$.

Solution 2. We present a different proof of (*).
We perform a sequence of small modification steps on the configuration of the circles in such a way that at the end they have no intersection at all (see Figure 6.1). We use two kinds of local changes to the structure of the orbits (see Figure 6.2):

- Type-1 step: An arc of a circle is moved over an arc of another circle; such a step creates or removes two intersections.
- Type-2 step: An arc of a circle is moved through the intersection of two other circles.


Figure 6.1



Figure 6.2

We assume that in every step only one circle is moved, and that this circle is moved over at most one arc or intersection point of other circles.

We will show that the parity of the number of orbits does not change in any step. As every circle becomes a separate orbit at the end of the procedure, this fact proves (*).

Consider what happens to the number of orbits when a Type-1 step is performed. The two intersection points are created or removed in a small neighbourhood. Denote some points of the two circles where they enter or leave this neighbourhood by $a, b, c$, and $d$ in this order around the neighbourhood; let $a$ and $b$ belong to one circle and let $c$ and $d$ belong to the other circle. The two circle arcs may have the same or opposite orientations. Moreover, the four end-points of the two arcs are connected by the other parts of the orbits. This can happen in two ways without intersection: either $a$ is connected to $d$ and $b$ is connected to $c$, or $a$ is connected to $b$ and $c$ is connected to $d$. Altogether we have four cases, as shown in Figure 7.


Figure 7
We can see that the number of orbits is changed by -2 or +2 in the leftmost case when the arcs have the same orientation, $a$ is connected to $d$, and $b$ is connected to $c$. In the other three cases the number of orbits is not changed. Hence, Type-1 steps do not change the parity of the number of orbits.

Now consider a Type-2 step. The three circles enclose a small, triangular region; by the step, this triangle is replaced by another triangle. Again, the modification of the orbits is done in some small neighbourhood; the structure does not change outside. Each side of the triangle shaped region can be convex or concave; the number of concave sides can be $0,1,2$ or 3 , so there are 4 possible arrangements of the orbits inside the neighbourhood, as shown in Figure 8.


Figure 8
Denote the points where the three circles enter or leave the neighbourhood by $a, b, c, d$, $e$, and $f$ in this order around the neighbourhood. As can be seen in Figure 8, there are only two essentially different cases; either $a, c, e$ are connected to $b, d, f$, respectively, or $a, c, e$ are connected to $f, b, d$, respectively. The step either preserves the set of connections or switches to the other arrangement. Obviously, in the earlier case the number of orbits is not changed; therefore we have to consider only the latter case.

The points $a, b, c, d, e$, and $f$ are connected by the orbits outside, without intersection. If $a$ was connected to $c$, say, then this orbit would isolate $b$, so this is impossible. Hence, each of $a, b, c, d, e$ and $f$ must be connected either to one of its neighbours or to the opposite point. If say $a$ is connected to $d$, then this orbit separates $b$ and $c$ from $e$ and $f$, therefore $b$ must be connected to $c$ and $e$ must be connected to $f$. Altogether there are only two cases and their reverses: either each point is connected to one of its neighbours or two opposite points are connected and the the remaining neigh boring pairs are connected to each other. See Figure 9.


Figure 9
We can see that if only neighbouring points are connected, then the number of orbits is changed by +2 or -2 . If two opposite points are connected ( $a$ and $d$ in the figure), then the orbits are re-arranged, but their number is unchanged. Hence, Type-2 steps also preserve the parity. This completes the proof of (*).

Solution 3. Like in the previous solutions, we do not need all circle pairs to intersect but we assume that the circles form a connected set. Denote by $\mathcal{C}$ and $\mathcal{P}$ the sets of circles and their intersection points, respectively.

The circles divide the plane into several simply connected, bounded regions and one unbounded region. Denote the set of these regions by $\mathcal{R}$. We say that an intersection point or a region is odd or even if it is contained inside an odd or even number of circles, respectively. Let $\mathcal{P}_{\text {odd }}$ and $\mathcal{R}_{\text {odd }}$ be the sets of odd intersection points and odd regions, respectively.

Claim.

$$
\begin{equation*}
\left|\mathcal{R}_{\text {odd }}\right|-\left|\mathcal{P}_{\text {odd }}\right| \equiv n \quad(\bmod 2) . \tag{1}
\end{equation*}
$$

Proof. For each circle $c \in \mathcal{C}$, denote by $R_{c}, P_{c}$, and $X_{c}$ the number of regions inside $c$, the number of intersection points inside $c$, and the number of circles intersecting $c$, respectively. The circles divide each other into several arcs; denote by $A_{c}$ the number of such arcs inside $c$. By double counting the regions and intersection points inside the circles we get

$$
\left|\mathcal{R}_{\text {odd }}\right| \equiv \sum_{c \in \mathcal{C}} R_{c} \quad(\bmod 2) \quad \text { and } \quad\left|\mathcal{P}_{\text {odd }}\right| \equiv \sum_{c \in \mathcal{C}} P_{c} \quad(\bmod 2) .
$$

For each circle $c$, apply EULER's polyhedron theorem to the (simply connected) regions in $c$. There are $2 X_{c}$ intersection points on $c$; they divide the circle into $2 X_{c}$ arcs. The polyhedron theorem yields $\left(R_{c}+1\right)+\left(P_{c}+2 X_{c}\right)=\left(A_{c}+2 X_{c}\right)+2$, considering the exterior of $c$ as a single region. Therefore,

$$
\begin{equation*}
R_{c}+P_{c}=A_{c}+1 \tag{2}
\end{equation*}
$$

Moreover, we have four arcs starting from every interior points inside $c$ and a single arc starting into the interior from each intersection point on the circle. By double-counting the end-points of the interior arcs we get $2 A_{c}=4 P_{c}+2 X_{c}$, so

$$
\begin{equation*}
A_{c}=2 P_{c}+X_{c} . \tag{3}
\end{equation*}
$$

The relations (2) and (3) together yield

$$
\begin{equation*}
R_{c}-P_{c}=X_{c}+1 \tag{4}
\end{equation*}
$$

By summing up (4) for all circles we obtain

$$
\sum_{c \in \mathcal{C}} R_{c}-\sum_{c \in \mathcal{C}} P_{c}=\sum_{c \in \mathcal{C}} X_{c}+|\mathcal{C}|,
$$

which yields

$$
\begin{equation*}
\left|\mathcal{R}_{\text {odd }}\right|-\left|\mathcal{P}_{\text {odd }}\right| \equiv \sum_{c \in \mathcal{C}} X_{c}+n \quad(\bmod 2) \tag{5}
\end{equation*}
$$

Notice that in $\sum_{c \in \mathcal{C}} X_{c}$ each intersecting circle pair is counted twice, i.e., for both circles in the pair, so

$$
\sum_{c \in \mathcal{C}} X_{c} \equiv 0 \quad(\bmod 2)
$$

which finishes the proof of the Claim.
Now insert the same small arcs at the intersections as in the first solution, and suppose that there is a single snail orbit $b$.

First we show that the odd regions are inside the curve $b$, while the even regions are outside. Take a region $r \in \mathcal{R}$ and a point $x$ in its interior, and draw a ray $y$, starting from $x$, that does not pass through any intersection point of the circles and is neither tangent to any of the circles. As is well-known, $x$ is inside the curve $b$ if and only if $y$ intersects $b$ an odd number of times (see Figure 10). Notice that if an arbitrary circle $c$ contains $x$ in its interior, then $c$ intersects $y$ at a single point; otherwise, if $x$ is outside $c$, then $c$ has 2 or 0 intersections with $y$. Therefore, $y$ intersects $b$ an odd number of times if and only if $x$ is contained in an odd number of circles, so if and only if $r$ is odd.


Figure 10
Now consider an intersection point $p$ of two circles $c_{1}$ and $c_{2}$ and a small neighbourhood around $p$. Suppose that $p$ is contained inside $k$ circles.

We have four regions that meet at $p$. Let $r_{1}$ be the region that lies outside both $c_{1}$ and $c_{2}$, let $r_{2}$ be the region that lies inside both $c_{1}$ and $c_{2}$, and let $r_{3}$ and $r_{4}$ be the two remaining regions, each lying inside exactly one of $c_{1}$ and $c_{2}$. The region $r_{1}$ is contained inside the same $k$ circles as $p$; the region $r_{2}$ is contained also by $c_{1}$ and $c_{2}$, so by $k+2$ circles in total; each of the regions $r_{3}$ and $r_{4}$ is contained inside $k+1$ circles. After the small arcs have been inserted at $p$, the regions $r_{1}$ and $r_{2}$ get connected, and the regions $r_{3}$ and $r_{4}$ remain separated at $p$ (see Figure 11). If $p$ is an odd point, then $r_{1}$ and $r_{2}$ are odd, so two odd regions are connected at $p$. Otherwise, if $p$ is even, then we have two even regions connected at $p$.


Figure 11


Figure 12

Consider the system of odd regions and their connections at the odd points as a graph. In this graph the odd regions are the vertices, and each odd point establishes an edge that connects two vertices (see Figure 12). As $b$ is a single closed curve, this graph is connected and contains no cycle, so the graph is a tree. Then the number of vertices must be by one greater than the number of edges, so

$$
\begin{equation*}
\left|\mathcal{R}_{\text {odd }}\right|-\left|\mathcal{P}_{\text {odd }}\right|=1 . \tag{9}
\end{equation*}
$$

The relations (1) and (9) together prove that $n$ must be odd.

Comment. For every odd $n$ there exists at least one configuration of $n$ circles with a single snail orbit. Figure 13 shows a possible configuration with 5 circles. In general, if a circle is rotated by $k \cdot \frac{360^{\circ}}{n}$ ( $k=1,2, \ldots, n-1$ ) around an interior point other than the centre, the circle and its rotated copies together provide a single snail orbit.


Figure 13

## Geometry

G1. The points $P$ and $Q$ are chosen on the side $B C$ of an acute-angled triangle $A B C$ so that $\angle P A B=\angle A C B$ and $\angle Q A C=\angle C B A$. The points $M$ and $N$ are taken on the rays $A P$ and $A Q$, respectively, so that $A P=P M$ and $A Q=Q N$. Prove that the lines $B M$ and $C N$ intersect on the circumcircle of the triangle $A B C$.
(Georgia)
Solution 1. Denote by $S$ the intersection point of the lines $B M$ and $C N$. Let moreover $\beta=\angle Q A C=\angle C B A$ and $\gamma=\angle P A B=\angle A C B$. From these equalities it follows that the triangles $A B P$ and $C A Q$ are similar (see Figure 1). Therefore we obtain

$$
\frac{B P}{P M}=\frac{B P}{P A}=\frac{A Q}{Q C}=\frac{N Q}{Q C}
$$

Moreover,

$$
\angle B P M=\beta+\gamma=\angle C Q N
$$

Hence the triangles $B P M$ and $N Q C$ are similar. This gives $\angle B M P=\angle N C Q$, so the triangles $B P M$ and $B S C$ are also similar. Thus we get

$$
\angle C S B=\angle B P M=\beta+\gamma=180^{\circ}-\angle B A C,
$$

which completes the solution.


Figure 1


Figure 2

Solution 2. As in the previous solution, denote by $S$ the intersection point of the lines $B M$ and $N C$. Let moreover the circumcircle of the triangle $A B C$ intersect the lines $A P$ and $A Q$ again at $K$ and $L$, respectively (see Figure 2).

Note that $\angle L B C=\angle L A C=\angle C B A$ and similarly $\angle K C B=\angle K A B=\angle B C A$. It implies that the lines $B L$ and $C K$ meet at a point $X$, being symmetric to the point $A$ with respect to the line $B C$. Since $A P=P M$ and $A Q=Q N$, it follows that $X$ lies on the line $M N$. Therefore, using Pascal's theorem for the hexagon $A L B S C K$, we infer that $S$ lies on the circumcircle of the triangle $A B C$, which finishes the proof.

Comment. Both solutions can be modified to obtain a more general result, with the equalities

$$
A P=P M \quad \text { and } \quad A Q=Q N
$$

replaced by

$$
\frac{A P}{P M}=\frac{Q N}{A Q}
$$

G2. Let $A B C$ be a triangle. The points $K, L$, and $M$ lie on the segments $B C, C A$, and $A B$, respectively, such that the lines $A K, B L$, and $C M$ intersect in a common point. Prove that it is possible to choose two of the triangles $A L M, B M K$, and $C K L$ whose inradii sum up to at least the inradius of the triangle $A B C$.
(Estonia)
Solution. Denote

$$
a=\frac{B K}{K C}, \quad b=\frac{C L}{L A}, \quad c=\frac{A M}{M B} .
$$

By Ceva's theorem, $a b c=1$, so we may, without loss of generality, assume that $a \geqslant 1$. Then at least one of the numbers $b$ or $c$ is not greater than 1 . Therefore at least one of the pairs $(a, b)$, $(b, c)$ has its first component not less than 1 and the second one not greater than 1 . Without loss of generality, assume that $1 \leqslant a$ and $b \leqslant 1$.

Therefore, we obtain $b c \leqslant 1$ and $1 \leqslant c a$, or equivalently

$$
\frac{A M}{M B} \leqslant \frac{L A}{C L} \quad \text { and } \quad \frac{M B}{A M} \leqslant \frac{B K}{K C}
$$

The first inequality implies that the line passing through $M$ and parallel to $B C$ intersects the segment $A L$ at a point $X$ (see Figure 1). Therefore the inradius of the triangle $A L M$ is not less than the inradius $r_{1}$ of triangle $A M X$.

Similarly, the line passing through $M$ and parallel to $A C$ intersects the segment $B K$ at a point $Y$, so the inradius of the triangle $B M K$ is not less than the inradius $r_{2}$ of the triangle $B M Y$. Thus, to complete our solution, it is enough to show that $r_{1}+r_{2} \geqslant r$, where $r$ is the inradius of the triangle $A B C$. We prove that in fact $r_{1}+r_{2}=r$.


Figure 1
Since $M X \| B C$, the dilation with centre $A$ that takes $M$ to $B$ takes the incircle of the triangle $A M X$ to the incircle of the triangle $A B C$. Therefore

$$
\frac{r_{1}}{r}=\frac{A M}{A B}, \quad \text { and similarly } \quad \frac{r_{2}}{r}=\frac{M B}{A B} .
$$

Adding these equalities gives $r_{1}+r_{2}=r$, as required.
Comment. Alternatively, one can use Desargues' theorem instead of Ceva's theorem, as follows: The lines $A B, B C, C A$ dissect the plane into seven regions. One of them is bounded, and amongst the other six, three are two-sided and three are three-sided. Now define the points $P=B C \cap L M$, $Q=C A \cap M K$, and $R=A B \cap K L$ (in the projective plane). By Desargues' theorem, the points $P$, $Q, R$ lie on a common line $\ell$. This line intersects only unbounded regions. If we now assume (without loss of generality) that $P, Q$ and $R$ lie on $\ell$ in that order, then one of the segments $P Q$ or $Q R$ lies inside a two-sided region. If, for example, this segment is $P Q$, then the triangles $A L M$ and $B M K$ will satisfy the statement of the problem for the same reason.

G3. Let $\Omega$ and $O$ be the circumcircle and the circumcentre of an acute-angled triangle $A B C$ with $A B>B C$. The angle bisector of $\angle A B C$ intersects $\Omega$ at $M \neq B$. Let $\Gamma$ be the circle with diameter $B M$. The angle bisectors of $\angle A O B$ and $\angle B O C$ intersect $\Gamma$ at points $P$ and $Q$, respectively. The point $R$ is chosen on the line $P Q$ so that $B R=M R$. Prove that $B R \| A C$. (Here we always assume that an angle bisector is a ray.)
(Russia)
Solution. Let $K$ be the midpoint of $B M$, i.e., the centre of $\Gamma$. Notice that $A B \neq B C$ implies $K \neq O$. Clearly, the lines $O M$ and $O K$ are the perpendicular bisectors of $A C$ and $B M$, respectively. Therefore, $R$ is the intersection point of $P Q$ and $O K$.

Let $N$ be the second point of intersection of $\Gamma$ with the line $O M$. Since $B M$ is a diameter of $\Gamma$, the lines $B N$ and $A C$ are both perpendicular to $O M$. Hence $B N \| A C$, and it suffices to prove that $B N$ passes through $R$. Our plan for doing this is to interpret the lines $B N, O K$, and $P Q$ as the radical axes of three appropriate circles.

Let $\omega$ be the circle with diameter $B O$. Since $\angle B N O=\angle B K O=90^{\circ}$, the points $N$ and $K$ lie on $\omega$.

Next we show that the points $O, K, P$, and $Q$ are concyclic. To this end, let $D$ and $E$ be the midpoints of $B C$ and $A B$, respectively. Clearly, $D$ and $E$ lie on the rays $O Q$ and $O P$, respectively. By our assumptions about the triangle $A B C$, the points $B, E, O, K$, and $D$ lie in this order on $\omega$. It follows that $\angle E O R=\angle E B K=\angle K B D=\angle K O D$, so the line $K O$ externally bisects the angle $P O Q$. Since the point $K$ is the centre of $\Gamma$, it also lies on the perpendicular bisector of $P Q$. So $K$ coincides with the midpoint of the $\operatorname{arc} P O Q$ of the circumcircle $\gamma$ of triangle $P O Q$.

Thus the lines $O K, B N$, and $P Q$ are pairwise radical axes of the circles $\omega, \gamma$, and $\Gamma$. Hence they are concurrent at $R$, as required.


G4. Consider a fixed circle $\Gamma$ with three fixed points $A, B$, and $C$ on it. Also, let us fix a real number $\lambda \in(0,1)$. For a variable point $P \notin\{A, B, C\}$ on $\Gamma$, let $M$ be the point on the segment $C P$ such that $C M=\lambda \cdot C P$. Let $Q$ be the second point of intersection of the circumcircles of the triangles $A M P$ and $B M C$. Prove that as $P$ varies, the point $Q$ lies on a fixed circle.

Solution 1. Throughout the solution, we denote by $\Varangle(a, b)$ the directed angle between the lines $a$ and $b$.

Let $D$ be the point on the segment $A B$ such that $B D=\lambda \cdot B A$. We will show that either $Q=D$, or $\Varangle(D Q, Q B)=\Varangle(A B, B C)$; this would mean that the point $Q$ varies over the constant circle through $D$ tangent to $B C$ at $B$, as required.

Denote the circumcircles of the triangles $A M P$ and $B M C$ by $\omega_{A}$ and $\omega_{B}$, respectively. The lines $A P, B C$, and $M Q$ are pairwise radical axes of the circles $\Gamma, \omega_{A}$, and $\omega_{B}$, thus either they are parallel, or they share a common point $X$.

Assume that these lines are parallel (see Figure 1). Then the segments $A P, Q M$, and $B C$ have a common perpendicular bisector; the reflection in this bisector maps the segment $C P$ to $B A$, and maps $M$ to $Q$. Therefore, in this case $Q$ lies on $A B$, and $B Q / A B=C M / C P=$ $B D / A B$; so we have $Q=D$.


Figure 1


Figure 2

Now assume that the lines $A P, Q M$, and $B C$ are concurrent at some point $X$ (see Figure 2). Notice that the points $A, B, Q$, and $X$ lie on a common circle $\Omega$ by Miquel's theorem applied to the triangle $X P C$. Let us denote by $Y$ the symmetric image of $X$ about the perpendicular bisector of $A B$. Clearly, $Y$ lies on $\Omega$, and the triangles $Y A B$ and $\triangle X B A$ are congruent. Moreover, the triangle $X P C$ is similar to the triangle $X B A$, so it is also similar to the triangle $Y A B$.

Next, the points $D$ and $M$ correspond to each other in similar triangles $Y A B$ and $X P C$, since $B D / B A=C M / C P=\lambda$. Moreover, the triangles $Y A B$ and $X P C$ are equi-oriented, so $\Varangle(M X, X P)=\Varangle(D Y, Y A)$. On the other hand, since the points $A, Q, X$, and $Y$ lie on $\Omega$, we have $\Varangle(Q Y, Y A)=\Varangle(M X, X P)$. Therefore, $\Varangle(Q Y, Y A)=\Varangle(D Y, Y A)$, so the points $Y, D$, and $Q$ are collinear.

Finally, we have $\Varangle(D Q, Q B)=\Varangle(Y Q, Q B)=\Varangle(Y A, A B)=\Varangle(A B, B X)=\Varangle(A B, B C)$, as desired.

Comment. In the original proposal, $\lambda$ was supposed to be an arbitrary real number distinct from 0 and 1 , and the point $M$ was defined by $\overrightarrow{C M}=\lambda \cdot \overrightarrow{C P}$. The Problem Selection Committee decided to add the restriction $\lambda \in(0,1)$ in order to avoid a large case distinction.
Solution 2. As in the previous solution, we introduce the radical centre $X=A P \cap B C \cap M Q$ of the circles $\omega_{A}, \omega_{B}$, and $\Gamma$. Next, we also notice that the points $A, Q, B$, and $X$ lie on a common circle $\Omega$.

If the point $P$ lies on the arc $B A C$ of $\Gamma$, then the point $X$ is outside $\Gamma$, thus the point $Q$ belongs to the ray $X M$, and therefore the points $P, A$, and $Q$ lie on the same side of $B C$. Otherwise, if $P$ lies on the arc $B C$ not containing $A$, then $X$ lies inside $\Gamma$, so $M$ and $Q$ lie on different sides of $B C$; thus again $Q$ and $A$ lie on the same side of $B C$. So, in each case the points $Q$ and $A$ lie on the same side of $B C$.


Figure 3
Now we prove that the ratio

$$
\frac{Q B}{\sin \angle Q B C}=\frac{Q B}{Q X} \cdot \frac{Q X}{\sin \angle Q B X}
$$

is constant. Since the points $A, Q, B$, and $X$ are concyclic, we have

$$
\frac{Q X}{\sin \angle Q B X}=\frac{A X}{\sin \angle A B C} .
$$

Next, since the points $B, Q, M$, and $C$ are concyclic, the triangles $X B Q$ and $X M C$ are similar, so

$$
\frac{Q B}{Q X}=\frac{C M}{C X}=\lambda \cdot \frac{C P}{C X} .
$$

Analogously, the triangles $X C P$ and $X A B$ are also similar, so

$$
\frac{C P}{C X}=\frac{A B}{A X}
$$

Therefore, we obtain

$$
\frac{Q B}{\sin \angle Q B C}=\lambda \cdot \frac{A B}{A X} \cdot \frac{A X}{\sin \angle A B C}=\lambda \cdot \frac{A B}{\sin \angle A B C}
$$

so this ratio is indeed constant. Thus the circle passing through $Q$ and tangent to $B C$ at $B$ is also constant, and $Q$ varies over this fixed circle.

Comment. It is not hard to guess that the desired circle should be tangent to $B C$ at $B$. Indeed, the second paragraph of this solution shows that this circle lies on one side of $B C$; on the other hand, in the limit case $P=B$, the point $Q$ also coincides with $B$.
Solution 3. Let us perform an inversion centred at $C$. Denote by $X^{\prime}$ the image of a point $X$ under this inversion.

The circle $\Gamma$ maps to the line $\Gamma^{\prime}$ passing through the constant points $A^{\prime}$ and $B^{\prime}$, and containing the variable point $P^{\prime}$. By the problem condition, the point $M$ varies over the circle $\gamma$ which is the homothetic image of $\Gamma$ with centre $C$ and coefficient $\lambda$. Thus $M^{\prime}$ varies over the constant line $\gamma^{\prime} \| A^{\prime} B^{\prime}$ which is the homothetic image of $A^{\prime} B^{\prime}$ with centre $C$ and coefficient $1 / \lambda$, and $M=\gamma^{\prime} \cap C P^{\prime}$. Next, the circumcircles $\omega_{A}$ and $\omega_{B}$ of the triangles $A M P$ and $B M C$ map to the circumcircle $\omega_{A}^{\prime}$ of the triangle $A^{\prime} M^{\prime} P^{\prime}$ and to the line $B^{\prime} M^{\prime}$, respectively; the point $Q$ thus maps to the second point of intersection of $B^{\prime} M^{\prime}$ with $\omega_{A}^{\prime}$ (see Figure 4).


Figure 4

Let $J$ be the (constant) common point of the lines $\gamma^{\prime}$ and $C A^{\prime}$, and let $\ell$ be the (constant) line through $J$ parallel to $C B^{\prime}$. Let $V$ be the common point of the lines $\ell$ and $B^{\prime} M^{\prime}$. Applying Pappus' theorem to the triples $\left(C, J, A^{\prime}\right)$ and $\left(V, B^{\prime}, M^{\prime}\right)$ we get that the points $C B^{\prime} \cap J V$, $J M^{\prime} \cap A^{\prime} B^{\prime}$, and $C M^{\prime} \cap A^{\prime} V$ are collinear. The first two of these points are ideal, hence so is the third, which means that $C M^{\prime} \| A^{\prime} V$.

Now we have $\Varangle\left(Q^{\prime} A^{\prime}, A^{\prime} P^{\prime}\right)=\Varangle\left(Q^{\prime} M^{\prime}, M^{\prime} P^{\prime}\right)=\angle\left(V M^{\prime}, A^{\prime} V\right)$, which means that the triangles $B^{\prime} Q^{\prime} A^{\prime}$ and $B^{\prime} A^{\prime} V$ are similar, and $\left(B^{\prime} A^{\prime}\right)^{2}=B^{\prime} Q^{\prime} \cdot B^{\prime} V$. Thus $Q^{\prime}$ is the image of $V$ under the second (fixed) inversion with centre $B^{\prime}$ and radius $B^{\prime} A^{\prime}$. Since $V$ varies over the constant line $\ell, Q^{\prime}$ varies over some constant circle $\Theta$. Thus, applying the first inversion back we get that $Q$ also varies over some fixed circle.

One should notice that this last circle is not a line; otherwise $\Theta$ would contain $C$, and thus $\ell$ would contain the image of $C$ under the second inversion. This is impossible, since $C B^{\prime} \| \ell$.

G5. Let $A B C D$ be a convex quadrilateral with $\angle B=\angle D=90^{\circ}$. Point $H$ is the foot of the perpendicular from $A$ to $B D$. The points $S$ and $T$ are chosen on the sides $A B$ and $A D$, respectively, in such a way that $H$ lies inside triangle $S C T$ and

$$
\angle S H C-\angle B S C=90^{\circ}, \quad \angle T H C-\angle D T C=90^{\circ} .
$$

Prove that the circumcircle of triangle $S H T$ is tangent to the line $B D$.

Solution. Let the line passing through $C$ and perpendicular to the line $S C$ intersect the line $A B$ at $Q$ (see Figure 1). Then

$$
\angle S Q C=90^{\circ}-\angle B S C=180^{\circ}-\angle S H C,
$$

which implies that the points $C, H, S$, and $Q$ lie on a common circle. Moreover, since $S Q$ is a diameter of this circle, we infer that the circumcentre $K$ of triangle $S H C$ lies on the line $A B$. Similarly, we prove that the circumcentre $L$ of triangle $C H T$ lies on the line $A D$.


Figure 1
In order to prove that the circumcircle of triangle $S H T$ is tangent to $B D$, it suffices to show that the perpendicular bisectors of $H S$ and $H T$ intersect on the line $A H$. However, these two perpendicular bisectors coincide with the angle bisectors of angles $A K H$ and $A L H$. Therefore, in order to complete the solution, it is enough (by the bisector theorem) to show that

$$
\begin{equation*}
\frac{A K}{K H}=\frac{A L}{L H} . \tag{1}
\end{equation*}
$$

We present two proofs of this equality.
First proof. Let the lines $K L$ and $H C$ intersect at $M$ (see Figure 2). Since $K H=K C$ and $L H=L C$, the points $H$ and $C$ are symmetric to each other with respect to the line $K L$. Therefore $M$ is the midpoint of $H C$. Denote by $O$ the circumcentre of quadrilateral $A B C D$. Then $O$ is the midpoint of $A C$. Therefore we have $O M \| A H$ and hence $O M \perp B D$. This together with the equality $O B=O D$ implies that $O M$ is the perpendicular bisector of $B D$ and therefore $B M=D M$.

Since $C M \perp K L$, the points $B, C, M$, and $K$ lie on a common circle with diameter $K C$. Similarly, the points $L, C, M$, and $D$ lie on a circle with diameter $L C$. Thus, using the sine law, we obtain

$$
\frac{A K}{A L}=\frac{\sin \angle A L K}{\sin \angle A K L}=\frac{D M}{C L} \cdot \frac{C K}{B M}=\frac{C K}{C L}=\frac{K H}{L H},
$$

which finishes the proof of (1).


Figure 2


Figure 3

Second proof. If the points $A, H$, and $C$ are collinear, then $A K=A L$ and $K H=L H$, so the equality (1) follows. Assume therefore that the points $A, H$, and $C$ do not lie in a line and consider the circle $\omega$ passing through them (see Figure 3). Since the quadrilateral $A B C D$ is cyclic,

$$
\angle B A C=\angle B D C=90^{\circ}-\angle A D H=\angle H A D .
$$

Let $N \neq A$ be the intersection point of the circle $\omega$ and the angle bisector of $\angle C A H$. Then $A N$ is also the angle bisector of $\angle B A D$. Since $H$ and $C$ are symmetric to each other with respect to the line $K L$ and $H N=N C$, it follows that both $N$ and the centre of $\omega$ lie on the line $K L$. This means that the circle $\omega$ is an Apollonius circle of the points $K$ and $L$. This immediately yields (1).

Comment. Either proof can be used to obtain the following generalised result:
Let $A B C D$ be a convex quadrilateral and let $H$ be a point in its interior with $\angle B A C=\angle D A H$. The points $S$ and $T$ are chosen on the sides $A B$ and $A D$, respectively, in such a way that $H$ lies inside triangle SCT and

$$
\angle S H C-\angle B S C=90^{\circ}, \quad \angle T H C-\angle D T C=90^{\circ} .
$$

Then the circumcentre of triangle SHT lies on the line AH (and moreover the circumcentre of triangle SCT lies on $A C$ ).

G6. Let $A B C$ be a fixed acute-angled triangle. Consider some points $E$ and $F$ lying on the sides $A C$ and $A B$, respectively, and let $M$ be the midpoint of $E F$. Let the perpendicular bisector of $E F$ intersect the line $B C$ at $K$, and let the perpendicular bisector of $M K$ intersect the lines $A C$ and $A B$ at $S$ and $T$, respectively. We call the pair $(E, F)$ interesting, if the quadrilateral $K S A T$ is cyclic.

Suppose that the pairs $\left(E_{1}, F_{1}\right)$ and $\left(E_{2}, F_{2}\right)$ are interesting. Prove that

$$
\frac{E_{1} E_{2}}{A B}=\frac{F_{1} F_{2}}{A C}
$$

Solution 1. For any interesting pair $(E, F)$, we will say that the corresponding triangle $E F K$ is also interesting.

Let $E F K$ be an interesting triangle. Firstly, we prove that $\angle K E F=\angle K F E=\angle A$, which also means that the circumcircle $\omega_{1}$ of the triangle $A E F$ is tangent to the lines $K E$ and $K F$.

Denote by $\omega$ the circle passing through the points $K, S, A$, and $T$. Let the line $A M$ intersect the line $S T$ and the circle $\omega$ (for the second time) at $N$ and $L$, respectively (see Figure 1).

Since $E F \| T S$ and $M$ is the midpoint of $E F, N$ is the midpoint of $S T$. Moreover, since $K$ and $M$ are symmetric to each other with respect to the line $S T$, we have $\angle K N S=\angle M N S=$ $\angle L N T$. Thus the points $K$ and $L$ are symmetric to each other with respect to the perpendicular bisector of $S T$. Therefore $K L \| S T$.

Let $G$ be the point symmetric to $K$ with respect to $N$. Then $G$ lies on the line $E F$, and we may assume that it lies on the ray $M F$. One has

$$
\angle K G E=\angle K N S=\angle S N M=\angle K L A=180^{\circ}-\angle K S A
$$

(if $K=L$, then the angle $K L A$ is understood to be the angle between $A L$ and the tangent to $\omega$ at $L$ ). This means that the points $K, G, E$, and $S$ are concyclic. Now, since $K S G T$ is a parallelogram, we obtain $\angle K E F=\angle K S G=180^{\circ}-\angle T K S=\angle A$. Since $K E=K F$, we also have $\angle K F E=\angle K E F=\angle A$.

After having proved this fact, one may finish the solution by different methods.


Figure 1


Figure 2

First method. We have just proved that all interesting triangles are similar to each other. This allows us to use the following lemma.

Lemma. Let $A B C$ be an arbitrary triangle. Choose two points $E_{1}$ and $E_{2}$ on the side $A C$, two points $F_{1}$ and $F_{2}$ on the side $A B$, and two points $K_{1}$ and $K_{2}$ on the side $B C$, in a way that the triangles $E_{1} F_{1} K_{1}$ and $E_{2} F_{2} K_{2}$ are similar. Then the six circumcircles of the triangles $A E_{i} F_{i}$, $B F_{i} K_{i}$, and $C E_{i} K_{i}(i=1,2)$ meet at a common point $Z$. Moreover, $Z$ is the centre of the spiral similarity that takes the triangle $E_{1} F_{1} K_{1}$ to the triangle $E_{2} F_{2} K_{2}$.
Proof. Firstly, notice that for each $i=1,2$, the circumcircles of the triangles $A E_{i} F_{i}, B F_{i} K_{i}$, and $C K_{i} E_{i}$ have a common point $Z_{i}$ by Miquel's theorem. Moreover, we have
$\Varangle\left(Z_{i} F_{i}, Z_{i} E_{i}\right)=\Varangle(A B, C A), \quad \Varangle\left(Z_{i} K_{i}, Z_{i} F_{i}\right)=\Varangle(B C, A B), \quad \Varangle\left(Z_{i} E_{i}, Z_{i} K_{i}\right)=\Varangle(C A, B C)$.
This yields that the points $Z_{1}$ and $Z_{2}$ correspond to each other in similar triangles $E_{1} F_{1} K_{1}$ and $E_{2} F_{2} K_{2}$. Thus, if they coincide, then this common point is indeed the desired centre of a spiral similarity.

Finally, in order to show that $Z_{1}=Z_{2}$, one may notice that $\Varangle\left(A B, A Z_{1}\right)=\Varangle\left(E_{1} F_{1}, E_{1} Z_{1}\right)=$ $\Varangle\left(E_{2} F_{2}, E_{2} Z_{2}\right)=\Varangle\left(A B, A Z_{2}\right)$ (see Figure 2). Similarly, one has $\Varangle\left(B C, B Z_{1}\right)=\Varangle\left(B C, B Z_{2}\right)$ and $\Varangle\left(C A, C Z_{1}\right)=\Varangle\left(C A, C Z_{2}\right)$. This yields $Z_{1}=Z_{2}$.

Now, let $P$ and $Q$ be the feet of the perpendiculars from $B$ and $C$ onto $A C$ and $A B$, respectively, and let $R$ be the midpoint of $B C$ (see Figure 3). Then $R$ is the circumcentre of the cyclic quadrilateral $B C P Q$. Thus we obtain $\angle A P Q=\angle B$ and $\angle R P C=\angle C$, which yields $\angle Q P R=\angle A$. Similarly, we show that $\angle P Q R=\angle A$. Thus, all interesting triangles are similar to the triangle $P Q R$.


Figure 3


Figure 4

Denote now by $Z$ the common point of the circumcircles of $A P Q, B Q R$, and $C P R$. Let $E_{1} F_{1} K_{1}$ and $E_{2} F_{2} K_{2}$ be two interesting triangles. By the lemma, $Z$ is the centre of any spiral similarity taking one of the triangles $E_{1} F_{1} K_{1}, E_{2} F_{2} K_{2}$, and $P Q R$ to some other of them. Therefore the triangles $Z E_{1} E_{2}$ and $Z F_{1} F_{2}$ are similar, as well as the triangles $Z E_{1} F_{1}$ and $Z P Q$. Hence

$$
\frac{E_{1} E_{2}}{F_{1} F_{2}}=\frac{Z E_{1}}{Z F_{1}}=\frac{Z P}{Z Q}
$$

Moreover, the equalities $\angle A Z Q=\angle A P Q=\angle A B C=180^{\circ}-\angle Q Z R$ show that the point $Z$ lies on the line $A R$ (see Figure 4). Therefore the triangles $A Z P$ and $A C R$ are similar, as well as the triangles $A Z Q$ and $A B R$. This yields

$$
\frac{Z P}{Z Q}=\frac{Z P}{R C} \cdot \frac{R B}{Z Q}=\frac{A Z}{A C} \cdot \frac{A B}{A Z}=\frac{A B}{A C}
$$

which completes the solution.

Second method. Now we will start from the fact that $\omega_{1}$ is tangent to the lines $K E$ and $K F$ (see Figure 5). We prove that if $(E, F)$ is an interesting pair, then

$$
\begin{equation*}
\frac{A E}{A B}+\frac{A F}{A C}=2 \cos \angle A \tag{1}
\end{equation*}
$$

Let $Y$ be the intersection point of the segments $B E$ and $C F$. The points $B, K$, and $C$ are collinear, hence applying PASCAL's theorem to the degenerated hexagon AFFYEE, we infer that $Y$ lies on the circle $\omega_{1}$.

Denote by $Z$ the second intersection point of the circumcircle of the triangle $B F Y$ with the line $B C$ (see Figure 6). By Miquel's theorem, the points $C, Z, Y$, and $E$ are concyclic. Therefore we obtain

$$
B F \cdot A B+C E \cdot A C=B Y \cdot B E+C Y \cdot C F=B Z \cdot B C+C Z \cdot B C=B C^{2}
$$

On the other hand, $B C^{2}=A B^{2}+A C^{2}-2 A B \cdot A C \cos \angle A$, by the cosine law. Hence

$$
(A B-A F) \cdot A B+(A C-A E) \cdot A C=A B^{2}+A C^{2}-2 A B \cdot A C \cos \angle A
$$

which simplifies to the desired equality (1).
Let now $\left(E_{1}, F_{1}\right)$ and $\left(E_{2}, F_{2}\right)$ be two interesting pairs of points. Then we get

$$
\frac{A E_{1}}{A B}+\frac{A F_{1}}{A C}=\frac{A E_{2}}{A B}+\frac{A F_{2}}{A C},
$$

which gives the desired result.


Figure 5


Figure 6

Third method. Again, we make use of the fact that all interesting triangles are similar (and equi-oriented). Let us put the picture onto a complex plane such that $A$ is at the origin, and identify each point with the corresponding complex number.

Let $E F K$ be any interesting triangle. The equalities $\angle K E F=\angle K F E=\angle A$ yield that the ratio $\nu=\frac{K-E}{F-E}$ is the same for all interesting triangles. This in turn means that the numbers $E$, $F$, and $K$ satisfy the linear equation

$$
\begin{equation*}
K=\mu E+\nu F, \quad \text { where } \quad \mu=1-\nu . \tag{2}
\end{equation*}
$$

Now let us choose the points $X$ and $Y$ on the rays $A B$ and $A C$, respectively, so that $\angle C X A=\angle A Y B=\angle A=\angle K E F$ (see Figure 7). Then each of the triangles $A X C$ and $Y A B$ is similar to any interesting triangle, which also means that

$$
\begin{equation*}
C=\mu A+\nu X=\nu X \quad \text { and } \quad B=\mu Y+\nu A=\mu Y \tag{3}
\end{equation*}
$$

Moreover, one has $X / Y=\overline{C / B}$.
Since the points $E, F$, and $K$ lie on $A C, A B$, and $B C$, respectively, one gets

$$
E=\rho Y, \quad F=\sigma X, \quad \text { and } \quad K=\lambda B+(1-\lambda) C
$$

for some real $\rho, \sigma$, and $\lambda$. In view of (3), the equation (2) now reads $\lambda B+(1-\lambda) C=K=$ $\mu E+\nu F=\rho B+\sigma C$, or

$$
(\lambda-\rho) B=(\sigma+\lambda-1) C .
$$

Since the nonzero complex numbers $B$ and $C$ have different arguments, the coefficients in the brackets vanish, so $\rho=\lambda$ and $\sigma=1-\lambda$. Therefore,

$$
\begin{equation*}
\frac{E}{Y}+\frac{F}{X}=\rho+\sigma=1 \tag{4}
\end{equation*}
$$

Now, if $\left(E_{1}, F_{1}\right)$ and $\left(E_{2}, F_{2}\right)$ are two distinct interesting pairs, one may apply (4) to both pairs. Subtracting, we get

$$
\frac{E_{1}-E_{2}}{Y}=\frac{F_{2}-F_{1}}{X}, \quad \text { so } \quad \frac{E_{1}-E_{2}}{F_{2}-F_{1}}=\frac{Y}{X}=\frac{\bar{B}}{\bar{C}}
$$

Taking absolute values provides the required result.


Figure 7

Comment 1. One may notice that the triangle $P Q R$ is also interesting.
Comment 2. In order to prove that $\angle K E F=\angle K F E=\angle A$, one may also use the following well-known fact:
Let $A E F$ be a triangle with $A E \neq A F$, and let $K$ be the common point of the symmedian taken from $A$ and the perpendicular bisector of $E F$. Then the lines $K E$ and $K F$ are tangent to the circumcircle $\omega_{1}$ of the triangle $A E F$.

In this case, however, one needs to deal with the case $A E=A F$ separately.

Solution 2. Let $(E, F)$ be an interesting pair. This time we prove that

$$
\begin{equation*}
\frac{A M}{A K}=\cos \angle A \tag{5}
\end{equation*}
$$

As in Solution 1, we introduce the circle $\omega$ passing through the points $K, S, A$, and $T$, together with the points $N$ and $L$ at which the line $A M$ intersect the line $S T$ and the circle $\omega$ for the second time, respectively. Let moreover $O$ be the centre of $\omega$ (see Figures 8 and 9). As in Solution 1, we note that $N$ is the midpoint of $S T$ and show that $K L \| S T$, which implies $\angle F A M=\angle E A K$.


Figure 8


Figure 9

Suppose now that $K \neq L$ (see Figure 8). Then $K L \| S T$, and consequently the lines $K M$ and $K L$ are perpendicular. It implies that the lines $L O$ and $K M$ meet at a point $X$ lying on the circle $\omega$. Since the lines $O N$ and $X M$ are both perpendicular to the line $S T$, they are parallel to each other, and hence $\angle L O N=\angle L X K=\angle M A K$. On the other hand, $\angle O L N=\angle M K A$, so we infer that triangles $N O L$ and $M A K$ are similar. This yields

$$
\frac{A M}{A K}=\frac{O N}{O L}=\frac{O N}{O T}=\cos \angle T O N=\cos \angle A
$$

If, on the other hand, $K=L$, then the points $A, M, N$, and $K$ lie on a common line, and this line is the perpendicular bisector of $S T$ (see Figure 9). This implies that $A K$ is a diameter of $\omega$, which yields $A M=2 O K-2 N K=2 O N$. So also in this case we obtain

$$
\frac{A M}{A K}=\frac{2 O N}{2 O T}=\cos \angle T O N=\cos \angle A
$$

Thus (5) is proved.
Let $P$ and $Q$ be the feet of the perpendiculars from $B$ and $C$ onto $A C$ and $A B$, respectively (see Figure 10). We claim that the point $M$ lies on the line $P Q$. Consider now the composition of the dilatation with factor $\cos \angle A$ and centre $A$, and the reflection with respect to the angle bisector of $\angle B A C$. This transformation is a similarity that takes $B, C$, and $K$ to $P, Q$, and $M$, respectively. Since $K$ lies on the line $B C$, the point $M$ lies on the line $P Q$.


Figure 10
Suppose that $E \neq P$. Then also $F \neq Q$, and by Menelaus' theorem, we obtain

$$
\frac{A Q}{F Q} \cdot \frac{F M}{E M} \cdot \frac{E P}{A P}=1
$$

Using the similarity of the triangles $A P Q$ and $A B C$, we infer that

$$
\frac{E P}{F Q}=\frac{A P}{A Q}=\frac{A B}{A C}, \quad \text { and hence } \quad \frac{E P}{A B}=\frac{F Q}{A C} .
$$

The last equality holds obviously also in case $E=P$, because then $F=Q$. Moreover, since the line $P Q$ intersects the segment $E F$, we infer that the point $E$ lies on the segment $A P$ if and only if the point $F$ lies outside of the segment $A Q$.

Let now $\left(E_{1}, F_{1}\right)$ and $\left(E_{2}, F_{2}\right)$ be two interesting pairs. Then we obtain

$$
\frac{E_{1} P}{A B}=\frac{F_{1} Q}{A C} \quad \text { and } \quad \frac{E_{2} P}{A B}=\frac{F_{2} Q}{A C} .
$$

If $P$ lies between the points $E_{1}$ and $E_{2}$, we add the equalities above, otherwise we subtract them. In any case we obtain

$$
\frac{E_{1} E_{2}}{A B}=\frac{F_{1} F_{2}}{A C},
$$

which completes the solution.

G7. Let $A B C$ be a triangle with circumcircle $\Omega$ and incentre $I$. Let the line passing through $I$ and perpendicular to $C I$ intersect the segment $B C$ and the arc $B C$ (not containing $A$ ) of $\Omega$ at points $U$ and $V$, respectively. Let the line passing through $U$ and parallel to $A I$ intersect $A V$ at $X$, and let the line passing through $V$ and parallel to $A I$ intersect $A B$ at $Y$. Let $W$ and $Z$ be the midpoints of $A X$ and $B C$, respectively. Prove that if the points $I, X$, and $Y$ are collinear, then the points $I, W$, and $Z$ are also collinear.
(U.S.A.)

Solution 1. We start with some general observations. Set $\alpha=\angle A / 2, \beta=\angle B / 2, \gamma=\angle C / 2$. Then obviously $\alpha+\beta+\gamma=90^{\circ}$. Since $\angle U I C=90^{\circ}$, we obtain $\angle I U C=\alpha+\beta$. Therefore $\angle B I V=\angle I U C-\angle I B C=\alpha=\angle B A I=\angle B Y V$, which implies that the points $B, Y, I$, and $V$ lie on a common circle (see Figure 1).

Assume now that the points $I, X$ and $Y$ are collinear. We prove that $\angle Y I A=90^{\circ}$.
Let the line $X U$ intersect $A B$ at $N$. Since the lines $A I, U X$, and $V Y$ are parallel, we get

$$
\frac{N X}{A I}=\frac{Y N}{Y A}=\frac{V U}{V I}=\frac{X U}{A I}
$$

implying $N X=X U$. Moreover, $\angle B I U=\alpha=\angle B N U$. This implies that the quadrilateral BUIN is cyclic, and since $B I$ is the angle bisector of $\angle U B N$, we infer that $N I=U I$. Thus in the isosceles triangle $N I U$, the point $X$ is the midpoint of the base $N U$. This gives $\angle I X N=90^{\circ}$, i.e., $\angle Y I A=90^{\circ}$.


Figure 1
Let $S$ be the midpoint of the segment $V C$. Let moreover $T$ be the intersection point of the lines $A X$ and $S I$, and set $x=\angle B A V=\angle B C V$. Since $\angle C I A=90^{\circ}+\beta$ and $S I=S C$, we obtain

$$
\angle T I A=180^{\circ}-\angle A I S=90^{\circ}-\beta-\angle C I S=90^{\circ}-\beta-\gamma-x=\alpha-x=\angle T A I,
$$

which implies that $T I=T A$. Therefore, since $\angle X I A=90^{\circ}$, the point $T$ is the midpoint of $A X$, i.e., $T=W$.

To complete our solution, it remains to show that the intersection point of the lines $I S$ and $B C$ coincide with the midpoint of the segment $B C$. But since $S$ is the midpoint of the segment $V C$, it suffices to show that the lines $B V$ and $I S$ are parallel.

Since the quadrilateral $B Y I V$ is cyclic, $\angle V B I=\angle V Y I=\angle Y I A=90^{\circ}$. This implies that $B V$ is the external angle bisector of the angle $A B C$, which yields $\angle V A C=\angle V C A$. Therefore $2 \alpha-x=2 \gamma+x$, which gives $\alpha=\gamma+x$. Hence $\angle S C I=\alpha$, so $\angle V S I=2 \alpha$.

On the other hand, $\angle B V C=180^{\circ}-\angle B A C=180^{\circ}-2 \alpha$, which implies that the lines $B V$ and $I S$ are parallel. This completes the solution.

Solution 2. As in Solution 1, we first prove that the points $B, Y, I, V$ lie on a common circle and $\angle Y I A=90^{\circ}$. The remaining part of the solution is based on the following lemma, which holds true for any triangle $A B C$, not necessarily with the property that $I, X, Y$ are collinear. Lemma. Let $A B C$ be the triangle inscribed in a circle $\Gamma$ and let $I$ be its incentre. Assume that the line passing through $I$ and perpendicular to the line $A I$ intersects the side $A B$ at the point $Y$. Let the circumcircle of the triangle $B Y I$ intersect the circle $\Gamma$ for the second time at $V$, and let the excircle of the triangle $A B C$ opposite to the vertex $A$ be tangent to the side $B C$ at $E$. Then

$$
\angle B A V=\angle C A E .
$$

Proof. Let $\rho$ be the composition of the inversion with centre $A$ and radius $\sqrt{A B \cdot A C}$, and the symmetry with respect to $A I$. Clearly, $\rho$ interchanges $B$ and $C$.

Let $J$ be the excentre of the triangle $A B C$ opposite to $A$ (see Figure 2). Then we have $\angle J A C=\angle B A I$ and $\angle J C A=90^{\circ}+\gamma=\angle B I A$, so the triangles $A C J$ and $A I B$ are similar, and therefore $A B \cdot A C=A I \cdot A J$. This means that $\rho$ interchanges $I$ and $J$. Moreover, since $Y$ lies on $A B$ and $\angle A I Y=90^{\circ}$, the point $Y^{\prime}=\rho(Y)$ lies on $A C$, and $\angle J Y^{\prime} A=90^{\circ}$. Thus $\rho$ maps the circumcircle $\gamma$ of the triangle $B Y I$ to a circle $\gamma^{\prime}$ with diameter $J C$.

Finally, since $V$ lies on both $\Gamma$ and $\gamma$, the point $V^{\prime}=\rho(V)$ lies on the line $\rho(\Gamma)=A B$ as well as on $\gamma^{\prime}$, which in turn means that $V^{\prime}=E$. This implies the desired result.


Figure 2


Figure 3

Now we turn to the solution of the problem.
Assume that the incircle $\omega_{1}$ of the triangle $A B C$ is tangent to $B C$ at $D$, and let the excircle $\omega_{2}$ of the triangle $A B C$ opposite to the vertex $A$ touch the side $B C$ at $E$ (see Figure 3). The homothety with centre $A$ that takes $\omega_{2}$ to $\omega_{1}$ takes the point $E$ to some point $F$, and the
tangent to $\omega_{1}$ at $F$ is parallel to $B C$. Therefore $D F$ is a diameter of $\omega_{1}$. Moreover, $Z$ is the midpoint of $D E$. This implies that the lines $I Z$ and $F E$ are parallel.

Let $K=Y I \cap A E$. Since $\angle Y I A=90^{\circ}$, the lemma yields that $I$ is the midpoint of $X K$. This implies that the segments $I W$ and $A K$ are parallel. Therefore, the points $W, I$ and $Z$ are collinear.

Comment 1. The properties $\angle Y I A=90^{\circ}$ and $V A=V C$ can be established in various ways. The main difficulty of the problem seems to find out how to use these properties in connection to the points $W$ and $Z$.

In Solution 2 this principal part is more or less covered by the lemma, for which we have presented a direct proof. On the other hand, this lemma appears to be a combination of two well-known facts; let us formulate them in terms of the lemma statement.

Let the line $I Y$ intersect $A C$ at $P$ (see Figure 4). The first fact states that the circumcircle $\omega$ of the triangle $V Y P$ is tangent to the segments $A B$ and $A C$, as well as to the circle $\Gamma$. The second fact states that for such a circle, the angles $B A V$ and $C A E$ are equal.

The awareness of this lemma may help a lot in solving this problem; so the Jury might also consider a variation of the proposed problem, for which the lemma does not seem to be useful; see Comment 3.


Comment 2. The proposed problem stated the equivalence: the point $I$ lies on the line $X Y$ if and only if $I$ lies on the line $W Z$. Here we sketch the proof of the "if" part (see Figure 5).
As in Solution 2, let $B C$ touch the circles $\omega_{1}$ and $\omega_{2}$ at $D$ and $E$, respectively. Since $I Z \| A E$ and $W$ lies on $I Z$, the line $D X$ is also parallel to $A E$. Therefore, the triangles $X U P$ and $A I Q$ are similar. Moreover, the line $D X$ is symmetric to $A E$ with respect to $I$, so $I P=I Q$, where $P=U V \cap X D$ and $Q=U V \cap A E$. Thus we obtain

$$
\frac{U V}{V I}=\frac{U X}{I A}=\frac{U P}{I Q}=\frac{U P}{I P}
$$

So the pairs $I U$ and $P V$ are harmonic conjugates, and since $\angle U D I=90^{\circ}$, we get $\angle V D B=\angle B D X=$ $\angle B E A$. Therefore the point $V^{\prime}$ symmetric to $V$ with respect to the perpendicular bisector of $B C$ lies on the line $A E$. So we obtain $\angle B A V=\angle C A E$.

The rest can be obtained by simply reversing the arguments in Solution 2 . The points $B, V, I$, and $Y$ are concyclic. The lemma implies that $\angle Y I A=90^{\circ}$. Moreover, the points $B, U, I$, and $N$, where $N=U X \cap A B$, lie on a common circle, so $I N=I U$. Since $I Y \perp U N$, the point $X^{\prime}=I Y \cap U N$ is the midpoint of $U N$. But in the trapezoid $A Y V I$, the line $X U$ is parallel to the sides $A I$ and $Y V$, so $N X=U X^{\prime}$. This yields $X=X^{\prime}$.
The reasoning presented in Solution 1 can also be reversed, but it requires a lot of technicalities. Therefore the Problem Selection Committee proposes to consider only the "only if" part of the original proposal, which is still challenging enough.

Comment 3. The Jury might also consider the following variation of the proposed problem.
Let $A B C$ be a triangle with circumcircle $\Omega$ and incentre $I$. Let the line through I perpendicular to $C I$ intersect the segment $B C$ and the arc $B C$ (not containing $A$ ) of $\Omega$ at $U$ and $V$, respectively. Let the line through $U$ parallel to $A I$ intersect $A V$ at $X$. Prove that if the lines XI and AI are perpendicular, then the midpoint of the segment AC lies on the line XI (see Figure 6).


Figure 6


Figure 7

Since the solution contains the arguments used above, we only sketch it.
Let $N=X U \cap A B$ (see Figure 7). Then $\angle B N U=\angle B A I=\angle B I U$, so the points $B, U, I$, and $N$ lie on a common circle. Therefore $I U=I N$, and since $I X \perp N U$, it follows that $N X=X U$.
Now set $Y=X I \cap A B$. The equality $N X=X U$ implies that

$$
\frac{V X}{V A}=\frac{X U}{A I}=\frac{N X}{A I}=\frac{Y X}{Y I},
$$

and therefore $Y V \| A I$. Hence $\angle B Y V=\angle B A I=\angle B I V$, so the points $B, V, I, Y$ are concyclic. Next we have $I Y \perp Y V$, so $\angle I B V=90^{\circ}$. This implies that $B V$ is the external angle bisector of the angle $A B C$, which gives $\angle V A C=\angle V C A$.
So in order to show that $M=X I \cap A C$ is the midpoint of $A C$, it suffices to prove that $\angle V M C=90^{\circ}$. But this follows immediately from the observation that the points $V, C, M$, and $I$ are concyclic, as $\angle M I V=\angle Y B V=180^{\circ}-\angle A C V$.
The converse statement is also true, but its proof requires some technicalities as well.

## Number Theory

N1. Let $n \geqslant 2$ be an integer, and let $A_{n}$ be the set

$$
A_{n}=\left\{2^{n}-2^{k} \mid k \in \mathbb{Z}, 0 \leqslant k<n\right\} .
$$

Determine the largest positive integer that cannot be written as the sum of one or more (not necessarily distinct) elements of $A_{n}$.
(Serbia)
Answer. $(n-2) 2^{n}+1$.

## Solution 1.

Part I. First we show that every integer greater than $(n-2) 2^{n}+1$ can be represented as such a sum. This is achieved by induction on $n$.

For $n=2$, the set $A_{n}$ consists of the two elements 2 and 3 . Every positive integer $m$ except for 1 can be represented as the sum of elements of $A_{n}$ in this case: as $m=2+2+\cdots+2$ if $m$ is even, and as $m=3+2+2+\cdots+2$ if $m$ is odd.

Now consider some $n>2$, and take an integer $m>(n-2) 2^{n}+1$. If $m$ is even, then consider

$$
\frac{m}{2} \geqslant \frac{(n-2) 2^{n}+2}{2}=(n-2) 2^{n-1}+1>(n-3) 2^{n-1}+1 .
$$

By the induction hypothesis, there is a representation of the form

$$
\frac{m}{2}=\left(2^{n-1}-2^{k_{1}}\right)+\left(2^{n-1}-2^{k_{2}}\right)+\cdots+\left(2^{n-1}-2^{k_{r}}\right)
$$

for some $k_{i}$ with $0 \leqslant k_{i}<n-1$. It follows that

$$
m=\left(2^{n}-2^{k_{1}+1}\right)+\left(2^{n}-2^{k_{2}+1}\right)+\cdots+\left(2^{n}-2^{k_{r}+1}\right)
$$

giving us the desired representation as a sum of elements of $A_{n}$. If $m$ is odd, we consider

$$
\frac{m-\left(2^{n}-1\right)}{2}>\frac{(n-2) 2^{n}+1-\left(2^{n}-1\right)}{2}=(n-3) 2^{n-1}+1
$$

By the induction hypothesis, there is a representation of the form

$$
\frac{m-\left(2^{n}-1\right)}{2}=\left(2^{n-1}-2^{k_{1}}\right)+\left(2^{n-1}-2^{k_{2}}\right)+\cdots+\left(2^{n-1}-2^{k_{r}}\right)
$$

for some $k_{i}$ with $0 \leqslant k_{i}<n-1$. It follows that

$$
m=\left(2^{n}-2^{k_{1}+1}\right)+\left(2^{n}-2^{k_{2}+1}\right)+\cdots+\left(2^{n}-2^{k_{r}+1}\right)+\left(2^{n}-1\right)
$$

giving us the desired representation of $m$ once again.
Part II. It remains to show that there is no representation for $(n-2) 2^{n}+1$. Let $N$ be the smallest positive integer that satisfies $N \equiv 1\left(\bmod 2^{n}\right)$, and which can be represented as a sum of elements of $A_{n}$. Consider a representation of $N$, i.e.,

$$
\begin{equation*}
N=\left(2^{n}-2^{k_{1}}\right)+\left(2^{n}-2^{k_{2}}\right)+\cdots+\left(2^{n}-2^{k_{r}}\right) \tag{1}
\end{equation*}
$$

where $0 \leqslant k_{1}, k_{2}, \ldots, k_{r}<n$. Suppose first that two of the terms in the sum are the same, i.e., $k_{i}=k_{j}$ for some $i \neq j$. If $k_{i}=k_{j}=n-1$, then we can simply remove these two terms to get a representation for

$$
N-2\left(2^{n}-2^{n-1}\right)=N-2^{n}
$$

as a sum of elements of $A_{n}$, which contradicts our choice of $N$. If $k_{i}=k_{j}=k<n-1$, replace the two terms by $2^{n}-2^{k+1}$, which is also an element of $A_{n}$, to get a representation for

$$
N-2\left(2^{n}-2^{k}\right)+2^{n}-2^{k+1}=N-2^{n}
$$

This is a contradiction once again. Therefore, all $k_{i}$ have to be distinct, which means that

$$
2^{k_{1}}+2^{k_{2}}+\cdots+2^{k_{r}} \leqslant 2^{0}+2^{1}+2^{2}+\cdots+2^{n-1}=2^{n}-1 .
$$

On the other hand, taking (1) modulo $2^{n}$, we find

$$
2^{k_{1}}+2^{k_{2}}+\cdots+2^{k_{r}} \equiv-N \equiv-1 \quad\left(\bmod 2^{n}\right)
$$

Thus we must have $2^{k_{1}}+2^{k_{2}}+\cdots+2^{k_{r}}=2^{n}-1$, which is only possible if each element of $\{0,1, \ldots, n-1\}$ occurs as one of the $k_{i}$. This gives us

$$
N=n 2^{n}-\left(2^{0}+2^{1}+\cdots+2^{n-1}\right)=(n-1) 2^{n}+1
$$

In particular, this means that $(n-2) 2^{n}+1$ cannot be represented as a sum of elements of $A_{n}$.
Solution 2. The fact that $m=(n-2) 2^{n}+1$ cannot be represented as a sum of elements of $A_{n}$ can also be shown in other ways. We prove the following statement by induction on $n$ :
Claim. If $a, b$ are integers with $a \geqslant 0, b \geqslant 1$, and $a+b<n$, then $a 2^{n}+b$ cannot be written as a sum of elements of $A_{n}$.
Proof. The claim is clearly true for $n=2$ (since $a=0, b=1$ is the only possibility). For $n>2$, assume that there exist integers $a, b$ with $a \geqslant 0, b \geqslant 1$ and $a+b<n$ as well as elements $m_{1}, m_{2}, \ldots, m_{r}$ of $A_{n}$ such that

$$
a 2^{n}+b=m_{1}+m_{2}+\cdots+m_{r} .
$$

We can suppose, without loss of generality, that $m_{1} \geqslant m_{2} \geqslant \cdots \geqslant m_{r}$. Let $\ell$ be the largest index for which $m_{\ell}=2^{n}-1\left(\ell=0\right.$ if $\left.m_{1} \neq 2^{n}-1\right)$. Clearly, $\ell$ and $b$ must have the same parity. Now

$$
(a-\ell) 2^{n}+(b+\ell)=m_{\ell+1}+m_{\ell+2}+\cdots+m_{r}
$$

and thus

$$
(a-\ell) 2^{n-1}+\frac{b+\ell}{2}=\frac{m_{\ell+1}}{2}+\frac{m_{\ell+2}}{2}+\cdots+\frac{m_{r}}{2} .
$$

Note that $m_{\ell+1} / 2, m_{\ell+2} / 2, \ldots, m_{r} / 2$ are elements of $A_{n-1}$. Moreover, $a-\ell$ and $(b+\ell) / 2$ are integers, and $(b+\ell) / 2 \geqslant 1$. If $a-\ell$ was negative, then we would have

$$
a 2^{n}+b \geqslant \ell\left(2^{n}-1\right) \geqslant(a+1)\left(2^{n}-1\right)=a 2^{n}+2^{n}-a-1,
$$

thus $n \geqslant a+b+1 \geqslant 2^{n}$, which is impossible. So $a-\ell \geqslant 0$. By the induction hypothesis, we must have $a-\ell+\frac{b+\ell}{2} \geqslant n-1$, which gives us a contradiction, since

$$
a-\ell+\frac{b+\ell}{2} \leqslant a-\ell+b+\ell-1=a+b-1<n-1 .
$$

Considering the special case $a=n-2, b=1$ now completes the proof.

Solution 3. Denote by $B_{n}$ the set of all positive integers that can be written as a sum of elements of $A_{n}$. In this solution, we explicitly describe all the numbers in $B_{n}$ by an argument similar to the first solution.

For a positive integer $n$, we denote by $\sigma_{2}(n)$ the sum of its digits in the binary representation. Notice that every positive integer $m$ has a unique representation of the form $m=s 2^{n}-t$ with some positive integer $s$ and $0 \leqslant t \leqslant 2^{n}-1$.
Lemma. For any two integers $s \geqslant 1$ and $0 \leqslant t \leqslant 2^{n}-1$, the number $m=s 2^{n}-t$ belongs to $B_{n}$ if and only if $s \geqslant \sigma_{2}(t)$.
Proof. For $t=0$, the statement of the Lemma is obvious, since $m=2 s \cdot\left(2^{n}-2^{n-1}\right)$.
Now suppose that $t \geqslant 1$, and let

$$
t=2^{k_{1}}+\cdots+2^{k_{\sigma}} \quad\left(0 \leqslant k_{1}<\cdots<k_{\sigma} \leqslant n-1, \quad \sigma=\sigma_{2}(t)\right)
$$

be its binary expansion. If $s \geqslant \sigma$, then $m \in B_{n}$ since

$$
m=(s-\sigma) 2^{n}+\left(\sigma 2^{n}-t\right)=2(s-\sigma) \cdot\left(2^{n}-2^{n-1}\right)+\sum_{i=1}^{\sigma}\left(2^{n}-2^{k_{i}}\right)
$$

Assume now that there exist integers $s$ and $t$ with $1 \leqslant s<\sigma_{2}(t)$ and $0 \leqslant t \leqslant 2^{n}-1$ such that the number $m=s 2^{n}-t$ belongs to $B_{n}$. Among all such instances, choose the one for which $m$ is smallest, and let

$$
m=\sum_{i=1}^{d}\left(2^{n}-2^{\ell_{i}}\right) \quad\left(0 \leqslant \ell_{i} \leqslant n-1\right)
$$

be the corresponding representation. If all the $\ell_{i}$ 's are distinct, then $\sum_{i=1}^{d} 2^{\ell_{i}} \leqslant \sum_{j=0}^{n-1} 2^{j}=2^{n}-1$, so one has $s=d$ and $t=\sum_{i=1}^{d} 2^{\ell_{i}}$, whence $s=d=\sigma_{2}(t)$; this is impossible. Therefore, two of the $\ell_{i}$ 's must be equal, say $\ell_{d-1}=\ell_{d}$. Then $m \geqslant 2\left(2^{n}-2^{\ell_{d}}\right) \geqslant 2^{n}$, so $s \geqslant 2$.

Now we claim that the number $m^{\prime}=m-2^{n}=(s-1) 2^{n}-t$ also belongs to $B_{n}$, which contradicts the minimality assumption. Indeed, one has

$$
\left(2^{n}-2^{\ell_{d-1}}\right)+\left(2^{n}-2^{\ell_{d}}\right)=2\left(2^{n}-2^{\ell_{d}}\right)=2^{n}+\left(2^{n}-2^{\ell_{d}+1}\right),
$$

so

$$
m^{\prime}=\sum_{i=1}^{d-2}\left(2^{n}-2^{\ell_{i}}\right)+\left(2^{n}-2^{\ell_{d}+1}\right)
$$

is the desired representation of $m^{\prime}$ (if $\ell_{d}=n-1$, then the last summand is simply omitted). This contradiction finishes the proof.

By our lemma, the largest number $M$ which does not belong to $B_{n}$ must have the form

$$
m_{t}=\left(\sigma_{2}(t)-1\right) 2^{n}-t
$$

for some $t$ with $1 \leqslant t \leqslant 2^{n}-1$, so $M$ is just the largest of these numbers. For $t_{0}=2^{n}-1$ we have $m_{t_{0}}=(n-1) 2^{n}-\left(2^{n}-1\right)=(n-2) 2^{n}+1$; for every other value of $t$ one has $\sigma_{2}(t) \leqslant n-1$, thus $m_{t} \leqslant(\sigma(t)-1) 2^{n} \leqslant(n-2) 2^{n}<m_{t_{0}}$. This means that $M=m_{t_{0}}=(n-2) 2^{n}+1$.

N2. Determine all pairs $(x, y)$ of positive integers such that

$$
\begin{equation*}
\sqrt[3]{7 x^{2}-13 x y+7 y^{2}}=|x-y|+1 \tag{1}
\end{equation*}
$$

Answer. Either $(x, y)=(1,1)$ or $\{x, y\}=\left\{m^{3}+m^{2}-2 m-1, m^{3}+2 m^{2}-m-1\right\}$ for some positive integer $m \geqslant 2$.
Solution. Let $(x, y)$ be any pair of positive integers solving (1). We shall prove that it appears in the list displayed above. The converse assertion that all these pairs do actually satisfy (1) either may be checked directly by means of a somewhat laborious calculation, or it can be seen by going in reverse order through the displayed equations that follow.

In case $x=y$ the given equation reduces to $x^{2 / 3}=1$, which is equivalent to $x=1$, whereby he have found the first solution.

To find the solutions with $x \neq y$ we may assume $x>y$ due to symmetry. Then the integer $n=x-y$ is positive and (1) may be rewritten as

$$
\sqrt[3]{7(y+n)^{2}-13(y+n) y+7 y^{2}}=n+1
$$

Raising this to the third power and simplifying the result one obtains

$$
y^{2}+y n=n^{3}-4 n^{2}+3 n+1 .
$$

To complete the square on the left hand side, we multiply by 4 and add $n^{2}$, thus getting

$$
(2 y+n)^{2}=4 n^{3}-15 n^{2}+12 n+4=(n-2)^{2}(4 n+1) .
$$

This shows that the cases $n=1$ and $n=2$ are impossible, whence $n>2$, and $4 n+1$ is the square of the rational number $\frac{2 y+n}{n-2}$. Consequently, it has to be a perfect square, and, since it is odd as well, there has to exist some nonnegative integer $m$ such that $4 n+1=(2 m+1)^{2}$, i.e.

$$
n=m^{2}+m .
$$

Notice that $n>2$ entails $m \geqslant 2$. Substituting the value of $n$ just found into the previous displayed equation we arrive at

$$
\left(2 y+m^{2}+m\right)^{2}=\left(m^{2}+m-2\right)^{2}(2 m+1)^{2}=\left(2 m^{3}+3 m^{2}-3 m-2\right)^{2} .
$$

Extracting square roots and taking $2 m^{3}+3 m^{2}-3 m-2=(m-1)\left(2 m^{2}+5 m+2\right)>0$ into account we derive $2 y+m^{2}+m=2 m^{3}+3 m^{2}-3 m-2$, which in turn yields

$$
y=m^{3}+m^{2}-2 m-1 .
$$

Notice that $m \geqslant 2$ implies that $y=\left(m^{3}-1\right)+(m-2) m$ is indeed positive, as it should be. In view of $x=y+n=y+m^{2}+m$ it also follows that

$$
x=m^{3}+2 m^{2}-m-1,
$$

and that this integer is positive as well.
Comment. Alternatively one could ask to find all pairs $(x, y)$ of - not necessarily positive - integers solving (1). The answer to that question is a bit nicer than the answer above: the set of solutions are now described by

$$
\{x, y\}=\left\{m^{3}+m^{2}-2 m-1, m^{3}+2 m^{2}-m-1\right\},
$$

where $m$ varies through $\mathbb{Z}$. This may be shown using essentially the same arguments as above. We finally observe that the pair $(x, y)=(1,1)$, that appears to be sporadic above, corresponds to $m=-1$.

N3. A coin is called a Cape Town coin if its value is $1 / n$ for some positive integer $n$. Given a collection of Cape Town coins of total value at most $99+\frac{1}{2}$, prove that it is possible to split this collection into at most 100 groups each of total value at most 1.
(Luxembourg)
Solution. We will show that for every positive integer $N$ any collection of Cape Town coins of total value at most $N-\frac{1}{2}$ can be split into $N$ groups each of total value at most 1 . The problem statement is a particular case for $N=100$.

We start with some preparations. If several given coins together have a total value also of the form $\frac{1}{k}$ for a positive integer $k$, then we may merge them into one new coin. Clearly, if the resulting collection can be split in the required way then the initial collection can also be split.

After each such merging, the total number of coins decreases, thus at some moment we come to a situation when no more merging is possible. At this moment, for every even $k$ there is at most one coin of value $\frac{1}{k}$ (otherwise two such coins may be merged), and for every odd $k>1$ there are at most $k-1$ coins of value $\frac{1}{k}$ (otherwise $k$ such coins may also be merged).

Now, clearly, each coin of value 1 should form a single group; if there are $d$ such coins then we may remove them from the collection and replace $N$ by $N-d$. So from now on we may assume that there are no coins of value 1 .

Finally, we may split all the coins in the following way. For each $k=1,2, \ldots, N$ we put all the coins of values $\frac{1}{2 k-1}$ and $\frac{1}{2 k}$ into a group $G_{k}$; the total value of $G_{k}$ does not exceed

$$
(2 k-2) \cdot \frac{1}{2 k-1}+\frac{1}{2 k}<1 .
$$

It remains to distribute the "small" coins of values which are less than $\frac{1}{2 N}$; we will add them one by one. In each step, take any remaining small coin. The total value of coins in the groups at this moment is at most $N-\frac{1}{2}$, so there exists a group of total value at most $\frac{1}{N}\left(N-\frac{1}{2}\right)=1-\frac{1}{2 N}$; thus it is possible to put our small coin into this group. Acting so, we will finally distribute all the coins.

Comment 1. The algorithm may be modified, at least the step where one distributes the coins of values $\geqslant \frac{1}{2 N}$. One different way is to put into $G_{k}$ all the coins of values $\frac{1}{(2 k-1) 2^{s}}$ for all integer $s \geqslant 0$. One may easily see that their total value also does not exceed 1 .

Comment 2. The original proposal also contained another part, suggesting to show that a required splitting may be impossible if the total value of coins is at most 100 . There are many examples of such a collection, e.g. one may take 98 coins of value 1 , one coin of value $\frac{1}{2}$, two coins of value $\frac{1}{3}$, and four coins of value $\frac{1}{5}$.

The Problem Selection Committee thinks that this part is less suitable for the competition.

N4. Let $n>1$ be a given integer. Prove that infinitely many terms of the sequence $\left(a_{k}\right)_{k \geqslant 1}$, defined by

$$
a_{k}=\left\lfloor\frac{n^{k}}{k}\right\rfloor,
$$

are odd. (For a real number $x,\lfloor x\rfloor$ denotes the largest integer not exceeding $x$.)
(Hong Kong)
Solution 1. If $n$ is odd, let $k=n^{m}$ for $m=1,2, \ldots$. Then $a_{k}=n^{n^{m}-m}$, which is odd for each $m$.

Henceforth, assume that $n$ is even, say $n=2 t$ for some integer $t \geqslant 1$. Then, for any $m \geqslant 2$, the integer $n^{2^{m}}-2^{m}=2^{m}\left(2^{2^{m}-m} \cdot t^{2^{m}}-1\right)$ has an odd prime divisor $p$, since $2^{m}-m>1$. Then, for $k=p \cdot 2^{m}$, we have

$$
n^{k}=\left(n^{2^{m}}\right)^{p} \equiv\left(2^{m}\right)^{p}=\left(2^{p}\right)^{m} \equiv 2^{m}
$$

where the congruences are taken modulo $p$ (recall that $2^{p} \equiv 2(\bmod p)$, by Fermat's little theorem). Also, from $n^{k}-2^{m}<n^{k}<n^{k}+2^{m}(p-1)$, we see that the fraction $\frac{n^{k}}{k}$ lies strictly between the consecutive integers $\frac{n^{k}-2^{m}}{p \cdot 2^{m}}$ and $\frac{n^{k}+2^{m}(p-1)}{p \cdot 2^{m}}$, which gives

$$
\left\lfloor\frac{n^{k}}{k}\right\rfloor=\frac{n^{k}-2^{m}}{p \cdot 2^{m}} .
$$

We finally observe that $\frac{n^{k}-2^{m}}{p \cdot 2^{m}}=\frac{\frac{n^{k}}{2^{m}}-1}{p}$ is an odd integer, since the integer $\frac{n^{k}}{2^{m}}-1$ is odd (recall that $k>m$ ). Note that for different values of $m$, we get different values of $k$, due to the different powers of 2 in the prime factorisation of $k$.

Solution 2. Treat the (trivial) case when $n$ is odd as in Solution 1.
Now assume that $n$ is even and $n>2$. Let $p$ be a prime divisor of $n-1$.
Proceed by induction on $i$ to prove that $p^{i+1}$ is a divisor of $n^{p^{i}}-1$ for every $i \geqslant 0$. The case $i=0$ is true by the way in which $p$ is chosen. Suppose the result is true for some $i \geqslant 0$. The factorisation

$$
n^{p^{i+1}}-1=\left(n^{p^{i}}-1\right)\left[n^{p^{i}(p-1)}+n^{p^{i}(p-2)}+\cdots+n^{p^{i}}+1\right],
$$

together with the fact that each of the $p$ terms between the square brackets is congruent to 1 modulo $p$, implies that the result is also true for $i+1$.

Hence $\left\lfloor\frac{n^{p^{i}}}{p^{i}}\right\rfloor=\frac{n^{p^{i}}-1}{p^{i}}$, an odd integer for each $i \geqslant 1$.
Finally, we consider the case $n=2$. We observe that $3 \cdot 4^{i}$ is a divisor of $2^{3 \cdot 4^{i}}-4^{i}$ for every $i \geqslant 1$ : Trivially, $4^{i}$ is a divisor of $2^{3 \cdot 4^{i}}-4^{i}$, since $3 \cdot 4^{i}>2 i$. Furthermore, since $2^{3 \cdot 4^{i}}$ and $4^{i}$ are both congruent to 1 modulo 3, we have $3 \mid 2^{3 \cdot 4^{i}}-4^{i}$. Hence, $\left\lfloor\frac{2^{3 \cdot 4^{i}}}{3 \cdot 4^{i}}\right\rfloor=\frac{2^{3 \cdot 4^{i}}-4^{i}}{3 \cdot 4^{i}}=\frac{2^{3 \cdot 4^{i}-2 i}-1}{3}$, which is odd for every $i \geqslant 1$.

Comment. The case $n$ even and $n>2$ can also be solved by recursively defining the sequence $\left(k_{i}\right)_{i \geqslant 1}$ by $k_{1}=1$ and $k_{i+1}=n^{k_{i}}-1$ for $i \geqslant 1$. Then $\left(k_{i}\right)$ is strictly increasing and it follows (by induction on $i$ ) that $k_{i} \mid n^{k_{i}}-1$ for all $i \geqslant 1$, so the $k_{i}$ are as desired.

The case $n=2$ can also be solved as follows: Let $i \geqslant 2$. By Bertrand's postulate, there exists a prime number $p$ such that $2^{2^{i}-1}<p \cdot 2^{i}<2^{2^{i}}$. This gives

$$
\begin{equation*}
p \cdot 2^{i}<2^{2^{i}}<2 p \cdot 2^{i} . \tag{1}
\end{equation*}
$$

Also, we have that $p \cdot 2^{i}$ is a divisor of $2^{p \cdot 2^{i}}-2^{2^{i}}$, hence, using (1), we get that

$$
\left\lfloor\frac{2^{p \cdot 2^{i}}}{p \cdot 2^{i}}\right\rfloor=\frac{2^{p \cdot 2^{i}}-2^{2^{i}}+p \cdot 2^{i}}{p \cdot 2^{i}}=\frac{2^{p \cdot 2^{i}-i}-2^{2^{i}-i}+p}{p}
$$

which is an odd integer.
Solution 3. Treat the (trivial) case when $n$ is odd as in Solution 1.
Let $n$ be even, and let $p$ be a prime divisor of $n+1$. Define the sequence $\left(a_{i}\right)_{i \geqslant 1}$ by

$$
a_{i}=\min \left\{a \in \mathbb{Z}_{>0}: 2^{i} \text { divides } a p+1\right\}
$$

Recall that there exists $a$ with $1 \leqslant a<2^{i}$ such that $a p \equiv-1\left(\bmod 2^{i}\right)$, so each $a_{i}$ satisfies $1 \leqslant a_{i}<2^{i}$. This implies that $a_{i} p+1<p \cdot 2^{i}$. Also, $a_{i} \rightarrow \infty$ as $i \rightarrow \infty$, whence there are infinitely many $i$ such that $a_{i}<a_{i+1}$. From now on, we restrict ourselves only to these $i$.

Notice that $p$ is a divisor of $n^{p}+1$, which, in turn, divides $n^{p \cdot 2^{i}}-1$. It follows that $p \cdot 2^{i}$ is a divisor of $n^{p \cdot 2^{i}}-\left(a_{i} p+1\right)$, and we consequently see that the integer $\left\lfloor\frac{n^{p \cdot 2^{i}}}{p \cdot 2^{i}}\right\rfloor=\frac{n^{p \cdot 2^{i}}-\left(a_{i} p+1\right)}{p \cdot 2^{i}}$ is odd, since $2^{i+1}$ divides $n^{p \cdot 2^{i}}$, but not $a_{i} p+1$.

N5. Find all triples $(p, x, y)$ consisting of a prime number $p$ and two positive integers $x$ and $y$ such that $x^{p-1}+y$ and $x+y^{p-1}$ are both powers of $p$.
(Belgium)
Answer. $(p, x, y) \in\{(3,2,5),(3,5,2)\} \cup\left\{\left(2, n, 2^{k}-n\right) \mid 0<n<2^{k}\right\}$.
Solution 1. For $p=2$, clearly all pairs of two positive integers $x$ and $y$ whose sum is a power of 2 satisfy the condition. Thus we assume in the following that $p>2$, and we let $a$ and $b$ be positive integers such that $x^{p-1}+y=p^{a}$ and $x+y^{p-1}=p^{b}$. Assume further, without loss of generality, that $x \leqslant y$, so that $p^{a}=x^{p-1}+y \leqslant x+y^{p-1}=p^{b}$, which means that $a \leqslant b$ (and thus $\left.p^{a} \mid p^{b}\right)$.

Now we have

$$
p^{b}=y^{p-1}+x=\left(p^{a}-x^{p-1}\right)^{p-1}+x .
$$

We take this equation modulo $p^{a}$ and take into account that $p-1$ is even, which gives us

$$
0 \equiv x^{(p-1)^{2}}+x \quad\left(\bmod p^{a}\right)
$$

If $p \mid x$, then $p^{a} \mid x$, since $x^{(p-1)^{2}-1}+1$ is not divisible by $p$ in this case. However, this is impossible, since $x \leqslant x^{p-1}<p^{a}$. Thus we know that $p \nmid x$, which means that

$$
p^{a} \mid x^{(p-1)^{2}-1}+1=x^{p(p-2)}+1
$$

By Fermat's little theorem, $x^{(p-1)^{2}} \equiv 1(\bmod p)$, thus $p$ divides $x+1$. Let $p^{r}$ be the highest power of $p$ that divides $x+1$. By the binomial theorem, we have

$$
x^{p(p-2)}=\sum_{k=0}^{p(p-2)}\binom{p(p-2)}{k}(-1)^{p(p-2)-k}(x+1)^{k} .
$$

Except for the terms corresponding to $k=0, k=1$ and $k=2$, all terms in the sum are clearly divisible by $p^{3 r}$ and thus by $p^{r+2}$. The remaining terms are

$$
-\frac{p(p-2)\left(p^{2}-2 p-1\right)}{2}(x+1)^{2}
$$

which is divisible by $p^{2 r+1}$ and thus also by $p^{r+2}$,

$$
p(p-2)(x+1)
$$

which is divisible by $p^{r+1}$, but not $p^{r+2}$ by our choice of $r$, and the final term -1 corresponding to $k=0$. It follows that the highest power of $p$ that divides $x^{p(p-2)}+1$ is $p^{r+1}$.

On the other hand, we already know that $p^{a}$ divides $x^{p(p-2)}+1$, which means that $a \leqslant r+1$. Moreover,

$$
p^{r} \leqslant x+1 \leqslant x^{p-1}+y=p^{a} .
$$

Hence we either have $a=r$ or $a=r+1$.
If $a=r$, then $x=y=1$ needs to hold in the inequality above, which is impossible for $p>2$. Thus $a=r+1$. Now since $p^{r} \leqslant x+1$, we get

$$
x=\frac{x^{2}+x}{x+1} \leqslant \frac{x^{p-1}+y}{x+1}=\frac{p^{a}}{x+1} \leqslant \frac{p^{a}}{p^{r}}=p,
$$

so we must have $x=p-1$ for $p$ to divide $x+1$.
It follows that $r=1$ and $a=2$. If $p \geqslant 5$, we obtain

$$
p^{a}=x^{p-1}+y>(p-1)^{4}=\left(p^{2}-2 p+1\right)^{2}>(3 p)^{2}>p^{2}=p^{a}
$$

a contradiction. So the only case that remains is $p=3$, and indeed $x=2$ and $y=p^{a}-x^{p-1}=5$ satisfy the conditions.

Comment 1. In this solution, we are implicitly using a special case of the following lemma known as "lifting the exponent":
Lemma. Let $n$ be a positive integer, let $p$ be an odd prime, and let $v_{p}(m)$ denote the exponent of the highest power of $p$ that divides $m$.

If $x$ and $y$ are integers not divisible by $p$ such that $p \mid x-y$, then we have

$$
v_{p}\left(x^{n}-y^{n}\right)=v_{p}(x-y)+v_{p}(n)
$$

Likewise, if $x$ and $y$ are integers not divisible by $p$ such that $p \mid x+y$, then we have

$$
v_{p}\left(x^{n}+y^{n}\right)=v_{p}(x+y)+v_{p}(n) .
$$

Comment 2. There exist various ways of solving the problem involving the "lifting the exponent" lemma. Let us sketch another one.

The cases $x=y$ and $p \mid x$ are ruled out easily, so we assume that $p>2, x<y$, and $p \nmid x$. In this case we also have $p^{a}<p^{b}$ and $p \mid x+1$.

Now one has

$$
y^{p}-x^{p} \equiv y\left(y^{p-1}+x\right)-x\left(x^{p-1}+y\right) \equiv 0 \quad\left(\bmod p^{a}\right),
$$

so by the lemma mentioned above one has $p^{a-1} \mid y-x$ and hence $y=x+t p^{a-1}$ for some positive integer $t$. Thus one gets

$$
x\left(x^{p-2}+1\right)=x^{p-1}+x=\left(x^{p-1}+y\right)-(y-x)=p^{a-1}(p-t) .
$$

The factors on the left-hand side are coprime. So if $p \mid x$, then $x^{p-2}+1 \mid p-t$, which is impossible since $x<x^{p-2}+1$. Therefore, $p \nmid x$, and thus $x \mid p-t$. Since $p \mid x+1$, the only remaining case is $x=p-1, t=1$, and $y=p^{a-1}+p-1$. Now the solution can be completed in the same way as before.
Solution 2. Again, we can focus on the case that $p>2$. If $p \mid x$, then also $p \mid y$. In this case, let $p^{k}$ and $p^{\ell}$ be the highest powers of $p$ that divide $x$ and $y$ respectively, and assume without loss of generality that $k \leqslant \ell$. Then $p^{k}$ divides $x+y^{p-1}$ while $p^{k+1}$ does not, but $p^{k}<x+y^{p-1}$, which yields a contradiction. So $x$ and $y$ are not divisible by $p$. Fermat's little theorem yields $0 \equiv x^{p-1}+y \equiv 1+y(\bmod p)$, so $y \equiv-1(\bmod p)$ and for the same reason $x \equiv-1(\bmod p)$.

In particular, $x, y \geqslant p-1$ and thus $x^{p-1}+y \geqslant 2(p-1)>p$, so $x^{p-1}+y$ and $y^{p-1}+x$ are both at least equal to $p^{2}$. Now we have

$$
x^{p-1} \equiv-y \quad\left(\bmod p^{2}\right) \quad \text { and } \quad y^{p-1} \equiv-x \quad\left(\bmod p^{2}\right)
$$

These two congruences, together with the Euler-Fermat theorem, give us

$$
1 \equiv x^{p(p-1)} \equiv(-y)^{p} \equiv-y^{p} \equiv x y \quad\left(\bmod p^{2}\right)
$$

Since $x \equiv y \equiv-1(\bmod p), x-y$ is divisible by $p$, so $(x-y)^{2}$ is divisible by $p^{2}$. This means that

$$
(x+y)^{2}=(x-y)^{2}+4 x y \equiv 4 \quad\left(\bmod p^{2}\right)
$$

so $p^{2}$ divides $(x+y-2)(x+y+2)$. We already know that $x+y \equiv-2(\bmod p)$, so $x+y-2 \equiv$ $-4 \not \equiv 0(\bmod p)$. This means that $p^{2}$ divides $x+y+2$.

Using the same notation as in the first solution, we subtract the two original equations to obtain

$$
\begin{equation*}
p^{b}-p^{a}=y^{p-1}-x^{p-1}+x-y=(y-x)\left(y^{p-2}+y^{p-3} x+\cdots+x^{p-2}-1\right) . \tag{1}
\end{equation*}
$$

The second factor is symmetric in $x$ and $y$, so it can be written as a polynomial of the elementary symmetric polynomials $x+y$ and $x y$ with integer coefficients. In particular, its value modulo
$p^{2}$ is characterised by the two congruences $x y \equiv 1\left(\bmod p^{2}\right)$ and $x+y \equiv-2\left(\bmod p^{2}\right)$. Since both congruences are satisfied when $x=y=-1$, we must have

$$
y^{p-2}+y^{p-3} x+\cdots+x^{p-2}-1 \equiv(-1)^{p-2}+(-1)^{p-3}(-1)+\cdots+(-1)^{p-2}-1 \quad\left(\bmod p^{2}\right)
$$

which simplifies to $y^{p-2}+y^{p-3} x+\cdots+x^{p-2}-1 \equiv-p\left(\bmod p^{2}\right)$. Thus the second factor in (1) is divisible by $p$, but not $p^{2}$.

This means that $p^{a-1}$ has to divide the other factor $y-x$. It follows that

$$
0 \equiv x^{p-1}+y \equiv x^{p-1}+x \equiv x(x+1)\left(x^{p-3}-x^{p-4}+\cdots+1\right) \quad\left(\bmod p^{a-1}\right)
$$

Since $x \equiv-1(\bmod p)$, the last factor is $x^{p-3}-x^{p-4}+\cdots+1 \equiv p-2(\bmod p)$ and in particular not divisible by $p$. We infer that $p^{a-1} \mid x+1$ and continue as in the first solution.

Comment. Instead of reasoning by means of elementary symmetric polynomials, it is possible to provide a more direct argument as well. For odd $r,(x+1)^{2}$ divides $\left(x^{r}+1\right)^{2}$, and since $p$ divides $x+1$, we deduce that $p^{2}$ divides $\left(x^{r}+1\right)^{2}$. Together with the fact that $x y \equiv 1\left(\bmod p^{2}\right)$, we obtain

$$
0 \equiv y^{r}\left(x^{r}+1\right)^{2} \equiv x^{2 r} y^{r}+2 x^{r} y^{r}+y^{r} \equiv x^{r}+2+y^{r} \quad\left(\bmod p^{2}\right) .
$$

We apply this congruence with $r=p-2-2 k$ (where $0 \leqslant k<(p-2) / 2$ ) to find that

$$
x^{k} y^{p-2-k}+x^{p-2-k} y^{k} \equiv(x y)^{k}\left(x^{p-2-2 k}+y^{p-2-2 k}\right) \equiv 1^{k} \cdot(-2) \equiv-2 \quad\left(\bmod p^{2}\right) .
$$

Summing over all $k$ yields

$$
y^{p-2}+y^{p-3} x+\cdots+x^{p-2}-1 \equiv \frac{p-1}{2} \cdot(-2)-1 \equiv-p \quad\left(\bmod p^{2}\right)
$$

once again.

N6. Let $a_{1}<a_{2}<\cdots<a_{n}$ be pairwise coprime positive integers with $a_{1}$ being prime and $a_{1} \geqslant n+2$. On the segment $I=\left[0, a_{1} a_{2} \cdots a_{n}\right]$ of the real line, mark all integers that are divisible by at least one of the numbers $a_{1}, \ldots, a_{n}$. These points split $I$ into a number of smaller segments. Prove that the sum of the squares of the lengths of these segments is divisible by $a_{1}$.
(Serbia)
Solution 1. Let $A=a_{1} \cdots a_{n}$. Throughout the solution, all intervals will be nonempty and have integer end-points. For any interval $X$, the length of $X$ will be denoted by $|X|$.

Define the following two families of intervals:

$$
\begin{aligned}
\mathcal{S} & =\{[x, y]: x<y \text { are consecutive marked points }\} \\
\mathcal{T} & =\{[x, y]: x<y \text { are integers, } 0 \leqslant x \leqslant A-1, \text { and no point is marked in }(x, y)\}
\end{aligned}
$$

We are interested in computing $\sum_{X \in \mathcal{S}}|X|^{2}$ modulo $a_{1}$.
Note that the number $A$ is marked, so in the definition of $\mathcal{T}$ the condition $y \leqslant A$ is enforced without explicitly prescribing it.

Assign weights to the intervals in $\mathcal{T}$, depending only on their lengths. The weight of an arbitrary interval $Y \in \mathcal{T}$ will be $w(|Y|)$, where

$$
w(k)= \begin{cases}1 & \text { if } k=1 \\ 2 & \text { if } k \geqslant 2\end{cases}
$$

Consider an arbitrary interval $X \in \mathcal{S}$ and its sub-intervals $Y \in \mathcal{T}$. Clearly, $X$ has one sub-interval of length $|X|$, two sub-intervals of length $|X|-1$ and so on; in general $X$ has $|X|-d+1$ sub-intervals of length $d$ for every $d=1,2, \ldots,|X|$. The sum of the weights of the sub-intervals of $X$ is
$\sum_{Y \in \mathcal{T}, Y \subseteq X} w(|Y|)=\sum_{d=1}^{|X|}(|X|-d+1) \cdot w(d)=|X| \cdot 1+((|X|-1)+(|X|-2)+\cdots+1) \cdot 2=|X|^{2}$.
Since the intervals in $\mathcal{S}$ are non-overlapping, every interval $Y \in \mathcal{T}$ is a sub-interval of a single interval $X \in \mathcal{S}$. Therefore,

$$
\begin{equation*}
\sum_{X \in \mathcal{S}}|X|^{2}=\sum_{X \in \mathcal{S}}\left(\sum_{Y \in \mathcal{T}, Y \subseteq X} w(|Y|)\right)=\sum_{Y \in \mathcal{T}} w(|Y|) . \tag{1}
\end{equation*}
$$

For every $d=1,2, \ldots, a_{1}$, we count how many intervals in $\mathcal{T}$ are of length $d$. Notice that the multiples of $a_{1}$ are all marked, so the lengths of the intervals in $\mathcal{S}$ and $\mathcal{T}$ cannot exceed $a_{1}$. Let $x$ be an arbitrary integer with $0 \leqslant x \leqslant A-1$ and consider the interval $[x, x+d]$. Let $r_{1}$, $\ldots, r_{n}$ be the remainders of $x$ modulo $a_{1}, \ldots, a_{n}$, respectively. Since $a_{1}, \ldots, a_{n}$ are pairwise coprime, the number $x$ is uniquely identified by the sequence $\left(r_{1}, \ldots, r_{n}\right)$, due to the Chinese remainder theorem.

For every $i=1, \ldots, n$, the property that the interval $(x, x+d)$ does not contain any multiple of $a_{i}$ is equivalent with $r_{i}+d \leqslant a_{i}$, i.e. $r_{i} \in\left\{0,1, \ldots, a_{i}-d\right\}$, so there are $a_{i}-d+1$ choices for the number $r_{i}$ for each $i$. Therefore, the number of the remainder sequences $\left(r_{1}, \ldots, r_{n}\right)$ that satisfy $[x, x+d] \in \mathcal{T}$ is precisely $\left(a_{1}+1-d\right) \cdots\left(a_{n}+1-d\right)$. Denote this product by $f(d)$.

Now we can group the last sum in (1) by length of the intervals. As we have seen, for every $d=1, \ldots, a_{1}$ there are $f(d)$ intervals $Y \in \mathcal{T}$ with $|Y|=d$. Therefore, (1) can be continued as

$$
\begin{equation*}
\sum_{X \in \mathcal{S}}|X|^{2}=\sum_{Y \in \mathcal{T}} w(|Y|)=\sum_{d=1}^{a_{1}} f(d) \cdot w(d)=2 \sum_{d=1}^{a_{1}} f(d)-f(1) . \tag{2}
\end{equation*}
$$

Having the formula (2), the solution can be finished using the following well-known fact: Lemma. If $p$ is a prime, $F(x)$ is a polynomial with integer coefficients, and $\operatorname{deg} F \leqslant p-2$, then $\sum_{x=1}^{p} F(x)$ is divisible by $p$.
Proof. Obviously, it is sufficient to prove the lemma for monomials of the form $x^{k}$ with $k \leqslant p-2$. Apply induction on $k$. If $k=0$ then $F=1$, and the statement is trivial.

Let $1 \leqslant k \leqslant p-2$, and assume that the lemma is proved for all lower degrees. Then

$$
\begin{aligned}
0 & \equiv p^{k+1}=\sum_{x=1}^{p}\left(x^{k+1}-(x-1)^{k+1}\right)=\sum_{x=1}^{p}\left(\sum_{\ell=0}^{k}(-1)^{k-\ell}\binom{k+1}{\ell} x^{\ell}\right) \\
& =(k+1) \sum_{x=1}^{p} x^{k}+\sum_{\ell=0}^{k-1}(-1)^{k-\ell}\binom{k+1}{\ell} \sum_{x=1}^{p} x^{\ell} \equiv(k+1) \sum_{x=1}^{p} x^{k} \quad(\bmod p) .
\end{aligned}
$$

Since $0<k+1<p$, this proves $\sum_{x=1}^{p} x^{k} \equiv 0(\bmod p)$.
In (2), by applying the lemma to the polynomial $f$ and the prime $a_{1}$, we obtain that $\sum_{d=1}^{a_{1}} f(d)$ is divisible by $a_{1}$. The term $f(1)=a_{1} \cdots a_{n}$ is also divisible by $a_{1}$; these two facts together prove that $\sum_{X \in \mathcal{S}}|X|^{2}$ is divisible by $a_{1}$.

Comment 1. With suitable sets of weights, the same method can be used to sum up other expressions on the lengths of the segments. For example, $w(1)=1$ and $w(k)=6(k-1)$ for $k \geqslant 2$ can be used to compute $\sum_{X \in \mathcal{S}}|X|^{3}$ and to prove that this sum is divisible by $a_{1}$ if $a_{1}$ is a prime with $a_{1} \geqslant n+3$. See also Comment 2 after the second solution.

Solution 2. The conventions from the first paragraph of the first solution are still in force. We shall prove the following more general statement:
( $\boxplus$ ) Let $p$ denote a prime number, let $p=a_{1}<a_{2}<\cdots<a_{n}$ be $n$ pairwise coprime positive integers, and let $d$ be an integer with $1 \leqslant d \leqslant p-n$. Mark all integers that are divisible by at least one of the numbers $a_{1}, \ldots, a_{n}$ on the interval $I=\left[0, a_{1} a_{2} \cdots a_{n}\right]$ of the real line. These points split $I$ into a number of smaller segments, say of lengths $b_{1}, \ldots, b_{k}$. Then the sum $\sum_{i=1}^{k}\binom{b_{i}}{d}$ is divisible by $p$.

Applying ( $\boxplus$ ) to $d=1$ and $d=2$ and using the equation $x^{2}=2\binom{x}{2}+\binom{x}{1}$, one easily gets the statement of the problem.

To prove ( $\boxplus$ ) itself, we argue by induction on $n$. The base case $n=1$ follows from the known fact that the binomial coefficient $\binom{p}{d}$ is divisible by $p$ whenever $1 \leqslant d \leqslant p-1$.

Let us now assume that $n \geqslant 2$, and that the statement is known whenever $n-1$ rather than $n$ coprime integers are given together with some integer $d \in[1, p-n+1]$. Suppose that
the numbers $p=a_{1}<a_{2}<\cdots<a_{n}$ and $d$ are as above. Write $A^{\prime}=\prod_{i=1}^{n-1} a_{i}$ and $A=A^{\prime} a_{n}$. Mark the points on the real axis divisible by one of the numbers $a_{1}, \ldots, a_{n-1}$ green and those divisible by $a_{n}$ red. The green points divide $\left[0, A^{\prime}\right]$ into certain sub-intervals, say $J_{1}, J_{2}, \ldots$, and $J_{\ell}$.

To translate intervals we use the notation $[a, b]+m=[a+m, b+m]$ whenever $a, b, m \in \mathbb{Z}$.
For each $i \in\{1,2, \ldots, \ell\}$ let $\mathcal{F}_{i}$ be the family of intervals into which the red points partition the intervals $J_{i}, J_{i}+A^{\prime}, \ldots$, and $J_{i}+\left(a_{n}-1\right) A^{\prime}$. We are to prove that

$$
\sum_{i=1}^{\ell} \sum_{X \in \mathcal{F}_{i}}\binom{|X|}{d}
$$

is divisible by $p$.
Let us fix any index $i$ with $1 \leqslant i \leqslant \ell$ for a while. Since the numbers $A^{\prime}$ and $a_{n}$ are coprime by hypothesis, the numbers $0, A^{\prime}, \ldots,\left(a_{n}-1\right) A^{\prime}$ form a complete system of residues modulo $a_{n}$. Moreover, we have $\left|J_{i}\right| \leqslant p<a_{n}$, as in particular all multiples of $p$ are green. So each of the intervals $J_{i}, J_{i}+A^{\prime}, \ldots$, and $J_{i}+\left(a_{n}-1\right) A^{\prime}$ contains at most one red point. More precisely, for each $j \in\left\{1, \ldots,\left|J_{i}\right|-1\right\}$ there is exactly one amongst those intervals containing a red point splitting it into an interval of length $j$ followed by an interval of length $\left|J_{i}\right|-j$, while the remaining $a_{n}-\left|J_{i}\right|+1$ such intervals have no red points in their interiors. For these reasons

$$
\begin{aligned}
\sum_{X \in \mathcal{F}_{i}}\binom{|X|}{d} & =2\left(\binom{1}{d}+\cdots+\binom{\left|J_{i}\right|-1}{d}\right)+\left(a_{n}-\left|J_{i}\right|+1\right)\binom{\left|J_{i}\right|}{d} \\
& =2\binom{\left|J_{i}\right|}{d+1}+\left(a_{n}-d+1\right)\binom{\left|J_{i}\right|}{d}-(d+1)\binom{\left|J_{i}\right|}{d+1} \\
& =(1-d)\binom{\left|J_{i}\right|}{d+1}+\left(a_{n}-d+1\right)\binom{\left|J_{i}\right|}{d} .
\end{aligned}
$$

So it remains to prove that

$$
(1-d) \sum_{i=1}^{\ell}\binom{\left|J_{i}\right|}{d+1}+\left(a_{n}-d+1\right) \sum_{i=1}^{\ell}\binom{\left|J_{i}\right|}{d}
$$

is divisible by $p$. By the induction hypothesis, however, it is even true that both summands are divisible by $p$, for $1 \leqslant d<d+1 \leqslant p-(n-1)$. This completes the proof of ( $\boxplus$ ) and hence the solution of the problem.

Comment 2. The statement ( $\boxplus$ ) can also be proved by the method of the first solution, using the weights $w(x)=\binom{x-2}{d-2}$.

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N7. Let $c \geqslant 1$ be an integer. Define a sequence of positive integers by $a_{1}=c$ and

$$
a_{n+1}=a_{n}^{3}-4 c \cdot a_{n}^{2}+5 c^{2} \cdot a_{n}+c
$$

for all $n \geqslant 1$. Prove that for each integer $n \geqslant 2$ there exists a prime number $p$ dividing $a_{n}$ but none of the numbers $a_{1}, \ldots, a_{n-1}$.
(Austria)
Solution. Let us define $x_{0}=0$ and $x_{n}=a_{n} / c$ for all integers $n \geqslant 1$. It is easy to see that the sequence ( $x_{n}$ ) thus obtained obeys the recursive law

$$
\begin{equation*}
x_{n+1}=c^{2}\left(x_{n}^{3}-4 x_{n}^{2}+5 x_{n}\right)+1 \tag{1}
\end{equation*}
$$

for all integers $n \geqslant 0$. In particular, all of its terms are positive integers; notice that $x_{1}=1$ and $x_{2}=2 c^{2}+1$. Since

$$
\begin{equation*}
x_{n+1}=c^{2} x_{n}\left(x_{n}-2\right)^{2}+c^{2} x_{n}+1>x_{n} \tag{2}
\end{equation*}
$$

holds for all integers $n \geqslant 0$, it is also strictly increasing. Since $x_{n+1}$ is by (1) coprime to $c$ for any $n \geqslant 0$, it suffices to prove that for each $n \geqslant 2$ there exists a prime number $p$ dividing $x_{n}$ but none of the numbers $x_{1}, \ldots, x_{n-1}$. Let us begin by establishing three preliminary claims.

Claim 1. If $i \equiv j(\bmod m)$ holds for some integers $i, j \geqslant 0$ and $m \geqslant 1$, then $x_{i} \equiv x_{j}\left(\bmod x_{m}\right)$ holds as well.

Proof. Evidently, it suffices to show $x_{i+m} \equiv x_{i}\left(\bmod x_{m}\right)$ for all integers $i \geqslant 0$ and $m \geqslant 1$. For this purpose we may argue for fixed $m$ by induction on $i$ using $x_{0}=0$ in the base case $i=0$. Now, if we have $x_{i+m} \equiv x_{i}\left(\bmod x_{m}\right)$ for some integer $i$, then the recursive equation (1) yields

$$
x_{i+m+1} \equiv c^{2}\left(x_{i+m}^{3}-4 x_{i+m}^{2}+5 x_{i+m}\right)+1 \equiv c^{2}\left(x_{i}^{3}-4 x_{i}^{2}+5 x_{i}\right)+1 \equiv x_{i+1} \quad\left(\bmod x_{m}\right),
$$

which completes the induction.
Claim 2. If the integers $i, j \geqslant 2$ and $m \geqslant 1$ satisfy $i \equiv j(\bmod m)$, then $x_{i} \equiv x_{j}\left(\bmod x_{m}^{2}\right)$ holds as well.
Proof. Again it suffices to prove $x_{i+m} \equiv x_{i}\left(\bmod x_{m}^{2}\right)$ for all integers $i \geqslant 2$ and $m \geqslant 1$. As above, we proceed for fixed $m$ by induction on $i$. The induction step is again easy using (1), but this time the base case $i=2$ requires some calculation. Set $L=5 c^{2}$. By (1) we have $x_{m+1} \equiv L x_{m}+1\left(\bmod x_{m}^{2}\right)$, and hence

$$
\begin{aligned}
x_{m+1}^{3}-4 x_{m+1}^{2}+5 x_{m+1} & \equiv\left(L x_{m}+1\right)^{3}-4\left(L x_{m}+1\right)^{2}+5\left(L x_{m}+1\right) \\
& \equiv\left(3 L x_{m}+1\right)-4\left(2 L x_{m}+1\right)+5\left(L x_{m}+1\right) \equiv 2 \quad\left(\bmod x_{m}^{2}\right)
\end{aligned}
$$

which in turn gives indeed $x_{m+2} \equiv 2 c^{2}+1 \equiv x_{2}\left(\bmod x_{m}^{2}\right)$.
Claim 3. For each integer $n \geqslant 2$, we have $x_{n}>x_{1} \cdot x_{2} \cdots x_{n-2}$.
Proof. The cases $n=2$ and $n=3$ are clear. Arguing inductively, we assume now that the claim holds for some $n \geqslant 3$. Recall that $x_{2} \geqslant 3$, so by monotonicity and (2) we get $x_{n} \geqslant x_{3} \geqslant x_{2}\left(x_{2}-2\right)^{2}+x_{2}+1 \geqslant 7$. It follows that

$$
x_{n+1}>x_{n}^{3}-4 x_{n}^{2}+5 x_{n}>7 x_{n}^{2}-4 x_{n}^{2}>x_{n}^{2}>x_{n} x_{n-1},
$$

which by the induction hypothesis yields $x_{n+1}>x_{1} \cdot x_{2} \cdots x_{n-1}$, as desired.

Now we direct our attention to the problem itself: let any integer $n \geqslant 2$ be given. By Claim 3 there exists a prime number $p$ appearing with a higher exponent in the prime factorisation of $x_{n}$ than in the prime factorisation of $x_{1} \cdots x_{n-2}$. In particular, $p \mid x_{n}$, and it suffices to prove that $p$ divides none of $x_{1}, \ldots, x_{n-1}$.

Otherwise let $k \in\{1, \ldots, n-1\}$ be minimal such that $p$ divides $x_{k}$. Since $x_{n-1}$ and $x_{n}$ are coprime by (1) and $x_{1}=1$, we actually have $2 \leqslant k \leqslant n-2$. Write $n=q k+r$ with some integers $q \geqslant 0$ and $0 \leqslant r<k$. By Claim 1 we have $x_{n} \equiv x_{r}\left(\bmod x_{k}\right)$, whence $p \mid x_{r}$. Due to the minimality of $k$ this entails $r=0$, i.e. $k \mid n$.

Thus from Claim 2 we infer

$$
x_{n} \equiv x_{k} \quad\left(\bmod x_{k}^{2}\right) .
$$

Now let $\alpha \geqslant 1$ be maximal with the property $p^{\alpha} \mid x_{k}$. Then $x_{k}^{2}$ is divisible by $p^{\alpha+1}$ and by our choice of $p$ so is $x_{n}$. So by the previous congruence $x_{k}$ is a multiple of $p^{\alpha+1}$ as well, contrary to our choice of $\alpha$. This is the final contradiction concluding the solution.

N8. For every real number $x$, let $\|x\|$ denote the distance between $x$ and the nearest integer. Prove that for every pair $(a, b)$ of positive integers there exist an odd prime $p$ and a positive integer $k$ satisfying

$$
\begin{equation*}
\left\|\frac{a}{p^{k}}\right\|+\left\|\frac{b}{p^{k}}\right\|+\left\|\frac{a+b}{p^{k}}\right\|=1 \tag{1}
\end{equation*}
$$

(Hungary)
Solution. Notice first that $\left\lfloor x+\frac{1}{2}\right\rfloor$ is an integer nearest to $x$, so $\|x\|=\left\lfloor\left.\left\lfloor x+\frac{1}{2}\right\rfloor-x \right\rvert\,\right.$. Thus we have

$$
\begin{equation*}
\left\lfloor x+\frac{1}{2}\right\rfloor=x \pm\|x\| \text {. } \tag{2}
\end{equation*}
$$

For every rational number $r$ and every prime number $p$, denote by $v_{p}(r)$ the exponent of $p$ in the prime factorisation of $r$. Recall the notation $(2 n-1)!$ ! for the product of all odd positive integers not exceeding $2 n-1$, i.e., $(2 n-1)!!=1 \cdot 3 \cdots(2 n-1)$.
Lemma. For every positive integer $n$ and every odd prime $p$, we have

$$
v_{p}((2 n-1)!!)=\sum_{k=1}^{\infty}\left\lfloor\frac{n}{p^{k}}+\frac{1}{2}\right\rfloor .
$$

Proof. For every positive integer $k$, let us count the multiples of $p^{k}$ among the factors $1,3, \ldots$, $2 n-1$. If $\ell$ is an arbitrary integer, the number $(2 \ell-1) p^{k}$ is listed above if and only if

$$
0<(2 \ell-1) p^{k} \leqslant 2 n \quad \Longleftrightarrow \quad \frac{1}{2}<\ell \leqslant \frac{n}{p^{k}}+\frac{1}{2} \quad \Longleftrightarrow \quad 1 \leqslant \ell \leqslant\left\lfloor\frac{n}{p^{k}}+\frac{1}{2}\right\rfloor
$$

Hence, the number of multiples of $p^{k}$ among the factors is precisely $m_{k}=\left\lfloor\frac{n}{p^{k}}+\frac{1}{2}\right\rfloor$. Thus we obtain

$$
v_{p}((2 n-1)!!)=\sum_{i=1}^{n} v_{p}(2 i-1)=\sum_{i=1}^{n} \sum_{k=1}^{v_{p}(2 i-1)} 1=\sum_{k=1}^{\infty} \sum_{\ell=1}^{m_{k}} 1=\sum_{k=1}^{\infty}\left\lfloor\frac{n}{p^{k}}+\frac{1}{2}\right\rfloor .
$$

In order to prove the problem statement, consider the rational number

$$
N=\frac{(2 a+2 b-1)!!}{(2 a-1)!!\cdot(2 b-1)!!}=\frac{(2 a+1)(2 a+3) \cdots(2 a+2 b-1)}{1 \cdot 3 \cdots(2 b-1)}
$$

Obviously, $N>1$, so there exists a prime $p$ with $v_{p}(N)>0$. Since $N$ is a fraction of two odd numbers, $p$ is odd.

By our lemma,

$$
0<v_{p}(N)=\sum_{k=1}^{\infty}\left(\left\lfloor\frac{a+b}{p^{k}}+\frac{1}{2}\right\rfloor-\left\lfloor\frac{a}{p^{k}}+\frac{1}{2}\right\rfloor-\left\lfloor\frac{b}{p^{k}}+\frac{1}{2}\right\rfloor\right) .
$$

Therefore, there exists some positive integer $k$ such that the integer number

$$
d_{k}=\left\lfloor\frac{a+b}{p^{k}}+\frac{1}{2}\right\rfloor-\left\lfloor\frac{a}{p^{k}}+\frac{1}{2}\right\rfloor-\left\lfloor\frac{b}{p^{k}}+\frac{1}{2}\right\rfloor
$$

is positive, so $d_{k} \geqslant 1$. By (2) we have

$$
\begin{equation*}
1 \leqslant d_{k}=\frac{a+b}{p^{k}}-\frac{a}{p^{k}}-\frac{b}{p^{k}} \pm\left\|\frac{a+b}{p^{k}}\right\| \pm\left\|\frac{a}{p^{k}}\right\| \pm\left\|\frac{b}{p^{k}}\right\|= \pm\left\|\frac{a+b}{p^{k}}\right\| \pm\left\|\frac{a}{p^{k}}\right\| \pm\left\|\frac{b}{p^{k}}\right\| \tag{3}
\end{equation*}
$$

Since $\|x\|<\frac{1}{2}$ for every rational $x$ with odd denominator, the relation (3) can only be satisfied if all three signs on the right-hand side are positive and $d_{k}=1$. Thus we get

$$
\left\|\frac{a}{p^{k}}\right\|+\left\|\frac{b}{p^{k}}\right\|+\left\|\frac{a+b}{p^{k}}\right\|=d_{k}=1,
$$

as required.
Comment 1. There are various choices for the number $N$ in the solution. Here we sketch such a version.

Let $x$ and $y$ be two rational numbers with odd denominators. It is easy to see that the condition $\|x\|+\|y\|+\|x+y\|=1$ is satisfied if and only if

$$
\text { either }\{x\}<\frac{1}{2}, \quad\{y\}<\frac{1}{2}, \quad\{x+y\}>\frac{1}{2}, \quad \text { or } \quad\{x\}>\frac{1}{2}, \quad\{y\}>\frac{1}{2}, \quad\{x+y\}<\frac{1}{2},
$$

where $\{x\}$ denotes the fractional part of $x$.
In the context of our problem, the first condition seems easier to deal with. Also, one may notice that

$$
\begin{equation*}
\{x\}<\frac{1}{2} \Longleftrightarrow \varkappa(x)=0 \quad \text { and } \quad\{x\} \geqslant \frac{1}{2} \Longleftrightarrow \varkappa(x)=1 \tag{4}
\end{equation*}
$$

where

$$
\varkappa(x)=\lfloor 2 x\rfloor-2\lfloor x\rfloor .
$$

Now it is natural to consider the number

$$
M=\frac{\binom{2 a+2 b}{a+b}}{\binom{2 a}{a}\binom{2 b}{b}}
$$

since

$$
v_{p}(M)=\sum_{k=1}^{\infty}\left(\varkappa\left(\frac{2(a+b)}{p^{k}}\right)-\varkappa\left(\frac{2 a}{p^{k}}\right)-\varkappa\left(\frac{2 b}{p^{k}}\right)\right) .
$$

One may see that $M>1$, and that $v_{2}(M) \leqslant 0$. Thus, there exist an odd prime $p$ and a positive integer $k$ with

$$
\varkappa\left(\frac{2(a+b)}{p^{k}}\right)-\varkappa\left(\frac{2 a}{p^{k}}\right)-\varkappa\left(\frac{2 b}{p^{k}}\right)>0 .
$$

In view of (4), the last inequality yields

$$
\begin{equation*}
\left\{\frac{a}{p^{k}}\right\}<\frac{1}{2}, \quad\left\{\frac{b}{p^{k}}\right\}<\frac{1}{2}, \quad \text { and } \quad\left\{\frac{a+b}{p^{k}}\right\}>\frac{1}{2}, \tag{5}
\end{equation*}
$$

which is what we wanted to obtain.
Comment 2. Once one tries to prove the existence of suitable $p$ and $k$ satisfying (5), it seems somehow natural to suppose that $a \leqslant b$ and to add the restriction $p^{k}>a$. In this case the inequalities (5) can be rewritten as

$$
2 a<p^{k}, \quad 2 m p^{k}<2 b<(2 m+1) p^{k}, \quad \text { and } \quad(2 m+1) p^{k}<2(a+b)<(2 m+2) p^{k}
$$

for some positive integer $m$. This means exactly that one of the numbers $2 a+1,2 a+3, \ldots, 2 a+2 b-1$ is divisible by some number of the form $p^{k}$ which is greater than $2 a$.

Using more advanced techniques, one can show that such a number $p^{k}$ exists even with $k=1$. This was shown in 2004 by Laishram and Shorey; the methods used for this proof are elementary but still quite involved. In fact, their result generalises a theorem by Sylvester which states that for every pair of integers $(n, k)$ with $n \geqslant k \geqslant 1$, the product $(n+1)(n+2) \cdots(n+k)$ is divisible by some prime $p>k$. We would like to mention here that Sylvester's theorem itself does not seem to suffice for solving the problem.

# Shortlisted Problems with Solutions 

## $56^{\text {th }}$ <br> International Mathematical Olympiad

# Shortlisted Problems with Solutions 

$56^{\text {th }}$ International Mathematical Olympiad
Chiang Mai, Thailand, 4-16


# The shortlisted problems should be kept strictly confidential until IMO 2016. 

## Contributing Countries

The Organizing Committee and the Problem Selection Committee of IMO 2015 thank the following 53 countries for contributing 155 problem proposals:

Albania, Algeria, Armenia, Australia, Austria, Brazil, Bulgaria, Canada, Costa Rica, Croatia, Cyprus, Denmark, El Salvador, Estonia, Finland, France, Georgia, Germany, Greece, Hong Kong, Hungary, India, Iran, Ireland, Israel, Italy, Japan, Kazakhstan, Lithuania, Luxembourg, Montenegro, Morocco, Netherlands, Pakistan, Poland, Romania, Russia, Saudi Arabia, Serbia, Singapore, Slovakia, Slovenia, South Africa, South Korea, Sweden, Turkey, Turkmenistan, Taiwan, Tanzania, Ukraine, United Kingdom, U.S.A., Uzbekistan

## Problem Selection Committee



Dungjade Shiowattana, Ilya I. Bogdanov, Tirasan Khandhawit, Wittawat Kositwattanarerk, Géza Kós, Weerachai Neeranartvong, Nipun Pitimanaaree, Christian Reiher, Nat Sothanaphan, Warut Suksompong, Wuttisak Trongsiriwat, Wijit Yangjit

## Problems

## Algebra

A1. Suppose that a sequence $a_{1}, a_{2}, \ldots$ of positive real numbers satisfies

$$
a_{k+1} \geqslant \frac{k a_{k}}{a_{k}^{2}+(k-1)}
$$

for every positive integer $k$. Prove that $a_{1}+a_{2}+\cdots+a_{n} \geqslant n$ for every $n \geqslant 2$.
(Serbia)
A2. Determine all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ with the property that

$$
f(x-f(y))=f(f(x))-f(y)-1
$$

holds for all $x, y \in \mathbb{Z}$.
(Croatia)
A3. Let $n$ be a fixed positive integer. Find the maximum possible value of

$$
\sum_{1 \leqslant r<s \leqslant 2 n}(s-r-n) x_{r} x_{s},
$$

where $-1 \leqslant x_{i} \leqslant 1$ for all $i=1,2, \ldots, 2 n$.
A4. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ satisfying the equation

$$
f(x+f(x+y))+f(x y)=x+f(x+y)+y f(x)
$$

for all real numbers $x$ and $y$.
(Albania)
A5. Let $2 \mathbb{Z}+1$ denote the set of odd integers. Find all functions $f: \mathbb{Z} \rightarrow 2 \mathbb{Z}+1$ satisfying

$$
f(x+f(x)+y)+f(x-f(x)-y)=f(x+y)+f(x-y)
$$

for every $x, y \in \mathbb{Z}$.
A6. Let $n$ be a fixed integer with $n \geqslant 2$. We say that two polynomials $P$ and $Q$ with real coefficients are block-similar if for each $i \in\{1,2, \ldots, n\}$ the sequences

$$
\begin{aligned}
& P(2015 i), P(2015 i-1), \ldots, P(2015 i-2014) \quad \text { and } \\
& Q(2015 i), Q(2015 i-1), \ldots, Q(2015 i-2014)
\end{aligned}
$$

are permutations of each other.
(a) Prove that there exist distinct block-similar polynomials of degree $n+1$.
(b) Prove that there do not exist distinct block-similar polynomials of degree $n$.

## Combinatorics

C1. In Lineland there are $n \geqslant 1$ towns, arranged along a road running from left to right. Each town has a left bulldozer (put to the left of the town and facing left) and a right bulldozer (put to the right of the town and facing right). The sizes of the $2 n$ bulldozers are distinct. Every time when a right and a left bulldozer confront each other, the larger bulldozer pushes the smaller one off the road. On the other hand, the bulldozers are quite unprotected at their rears; so, if a bulldozer reaches the rear-end of another one, the first one pushes the second one off the road, regardless of their sizes.

Let $A$ and $B$ be two towns, with $B$ being to the right of $A$. We say that town $A$ can sweep town $B$ away if the right bulldozer of $A$ can move over to $B$ pushing off all bulldozers it meets. Similarly, $B$ can sweep $A$ away if the left bulldozer of $B$ can move to $A$ pushing off all bulldozers of all towns on its way.

Prove that there is exactly one town which cannot be swept away by any other one.
(Estonia)
C2. Let $\mathcal{V}$ be a finite set of points in the plane. We say that $\mathcal{V}$ is balanced if for any two distinct points $A, B \in \mathcal{V}$, there exists a point $C \in \mathcal{V}$ such that $A C=B C$. We say that $\mathcal{V}$ is center-free if for any distinct points $A, B, C \in \mathcal{V}$, there does not exist a point $P \in \mathcal{V}$ such that $P A=P B=P C$.
(a) Show that for all $n \geqslant 3$, there exists a balanced set consisting of $n$ points.
(b) For which $n \geqslant 3$ does there exist a balanced, center-free set consisting of $n$ points?
(Netherlands)
C3. For a finite set $A$ of positive integers, we call a partition of $A$ into two disjoint nonempty subsets $A_{1}$ and $A_{2}$ good if the least common multiple of the elements in $A_{1}$ is equal to the greatest common divisor of the elements in $A_{2}$. Determine the minimum value of $n$ such that there exists a set of $n$ positive integers with exactly 2015 good partitions.
(Ukraine)
$\mathbf{C 4}$. Let $n$ be a positive integer. Two players $A$ and $B$ play a game in which they take turns choosing positive integers $k \leqslant n$. The rules of the game are:
(i) A player cannot choose a number that has been chosen by either player on any previous turn.
(ii) A player cannot choose a number consecutive to any of those the player has already chosen on any previous turn.
(iii) The game is a draw if all numbers have been chosen; otherwise the player who cannot choose a number anymore loses the game.

The player $A$ takes the first turn. Determine the outcome of the game, assuming that both players use optimal strategies.
(Finland)

C5. Consider an infinite sequence $a_{1}, a_{2}, \ldots$ of positive integers with $a_{i} \leqslant 2015$ for all $i \geqslant 1$. Suppose that for any two distinct indices $i$ and $j$ we have $i+a_{i} \neq j+a_{j}$.

Prove that there exist two positive integers $b$ and $N$ such that

$$
\left|\sum_{i=m+1}^{n}\left(a_{i}-b\right)\right| \leqslant 1007^{2}
$$

whenever $n>m \geqslant N$.
(Australia)
C6. Let $S$ be a nonempty set of positive integers. We say that a positive integer $n$ is clean if it has a unique representation as a sum of an odd number of distinct elements from $S$. Prove that there exist infinitely many positive integers that are not clean.

C7. In a company of people some pairs are enemies. A group of people is called unsociable if the number of members in the group is odd and at least 3 , and it is possible to arrange all its members around a round table so that every two neighbors are enemies. Given that there are at most 2015 unsociable groups, prove that it is possible to partition the company into 11 parts so that no two enemies are in the same part.

## Geometry

G1. Let $A B C$ be an acute triangle with orthocenter $H$. Let $G$ be the point such that the quadrilateral $A B G H$ is a parallelogram. Let $I$ be the point on the line $G H$ such that $A C$ bisects $H I$. Suppose that the line $A C$ intersects the circumcircle of the triangle $G C I$ at $C$ and $J$. Prove that $I J=A H$.
(Australia)
G2. Let $A B C$ be a triangle inscribed into a circle $\Omega$ with center $O$. A circle $\Gamma$ with center $A$ meets the side $B C$ at points $D$ and $E$ such that $D$ lies between $B$ and $E$. Moreover, let $F$ and $G$ be the common points of $\Gamma$ and $\Omega$. We assume that $F$ lies on the arc $A B$ of $\Omega$ not containing $C$, and $G$ lies on the arc $A C$ of $\Omega$ not containing $B$. The circumcircles of the triangles $B D F$ and $C E G$ meet the sides $A B$ and $A C$ again at $K$ and $L$, respectively. Suppose that the lines $F K$ and $G L$ are distinct and intersect at $X$. Prove that the points $A, X$, and $O$ are collinear.
(Greece)
G3. Let $A B C$ be a triangle with $\angle C=90^{\circ}$, and let $H$ be the foot of the altitude from $C$. A point $D$ is chosen inside the triangle $C B H$ so that $C H$ bisects $A D$. Let $P$ be the intersection point of the lines $B D$ and $C H$. Let $\omega$ be the semicircle with diameter $B D$ that meets the segment $C B$ at an interior point. A line through $P$ is tangent to $\omega$ at $Q$. Prove that the lines $C Q$ and $A D$ meet on $\omega$.
(Georgia)
G4. Let $A B C$ be an acute triangle, and let $M$ be the midpoint of $A C$. A circle $\omega$ passing through $B$ and $M$ meets the sides $A B$ and $B C$ again at $P$ and $Q$, respectively. Let $T$ be the point such that the quadrilateral $B P T Q$ is a parallelogram. Suppose that $T$ lies on the circumcircle of the triangle $A B C$. Determine all possible values of $B T / B M$.
(Russia)
G5. Let $A B C$ be a triangle with $C A \neq C B$. Let $D, F$, and $G$ be the midpoints of the sides $A B, A C$, and $B C$, respectively. A circle $\Gamma$ passing through $C$ and tangent to $A B$ at $D$ meets the segments $A F$ and $B G$ at $H$ and $I$, respectively. The points $H^{\prime}$ and $I^{\prime}$ are symmetric to $H$ and $I$ about $F$ and $G$, respectively. The line $H^{\prime} I^{\prime}$ meets $C D$ and $F G$ at $Q$ and $M$, respectively. The line $C M$ meets $\Gamma$ again at $P$. Prove that $C Q=Q P$.
(El Salvador)
G6. Let $A B C$ be an acute triangle with $A B>A C$, and let $\Gamma$ be its circumcircle. Let $H$, $M$, and $F$ be the orthocenter of the triangle, the midpoint of $B C$, and the foot of the altitude from $A$, respectively. Let $Q$ and $K$ be the two points on $\Gamma$ that satisfy $\angle A Q H=90^{\circ}$ and $\angle Q K H=90^{\circ}$. Prove that the circumcircles of the triangles $K Q H$ and $K F M$ are tangent to each other.
(Ukraine)
G7. Let $A B C D$ be a convex quadrilateral, and let $P, Q, R$, and $S$ be points on the sides $A B, B C, C D$, and $D A$, respectively. Let the line segments $P R$ and $Q S$ meet at $O$. Suppose that each of the quadrilaterals $A P O S, B Q O P, C R O Q$, and $D S O R$ has an incircle. Prove that the lines $A C, P Q$, and $R S$ are either concurrent or parallel to each other.
(Bulgaria)
G8. A triangulation of a convex polygon $\Pi$ is a partitioning of $\Pi$ into triangles by diagonals having no common points other than the vertices of the polygon. We say that a triangulation is a Thaiangulation if all triangles in it have the same area.

Prove that any two different Thaiangulations of a convex polygon $\Pi$ differ by exactly two triangles. (In other words, prove that it is possible to replace one pair of triangles in the first Thaiangulation with a different pair of triangles so as to obtain the second Thaiangulation.)
(Bulgaria)

## Number Theory

N1. Determine all positive integers $M$ for which the sequence $a_{0}, a_{1}, a_{2}, \ldots$, defined by $a_{0}=\frac{2 M+1}{2}$ and $a_{k+1}=a_{k}\left\lfloor a_{k}\right\rfloor$ for $k=0,1,2, \ldots$, contains at least one integer term.
(Luxembourg)
N2. Let $a$ and $b$ be positive integers such that $a!b!$ is a multiple of $a!+b!$. Prove that $3 a \geqslant 2 b+2$.
(United Kingdom)
N3. Let $m$ and $n$ be positive integers such that $m>n$. Define $x_{k}=(m+k) /(n+k)$ for $k=$ $1,2, \ldots, n+1$. Prove that if all the numbers $x_{1}, x_{2}, \ldots, x_{n+1}$ are integers, then $x_{1} x_{2} \cdots x_{n+1}-1$ is divisible by an odd prime.
(Austria)
N4. Suppose that $a_{0}, a_{1}, \ldots$ and $b_{0}, b_{1}, \ldots$ are two sequences of positive integers satisfying $a_{0}, b_{0} \geqslant 2$ and

$$
a_{n+1}=\operatorname{gcd}\left(a_{n}, b_{n}\right)+1, \quad b_{n+1}=\operatorname{lcm}\left(a_{n}, b_{n}\right)-1
$$

for all $n \geqslant 0$. Prove that the sequence $\left(a_{n}\right)$ is eventually periodic; in other words, there exist integers $N \geqslant 0$ and $t>0$ such that $a_{n+t}=a_{n}$ for all $n \geqslant N$.
(France)
N5. Determine all triples $(a, b, c)$ of positive integers for which $a b-c, b c-a$, and $c a-b$ are powers of 2 .

Explanation: A power of 2 is an integer of the form $2^{n}$, where $n$ denotes some nonnegative integer.
(Serbia)
N6. Let $\mathbb{Z}_{>0}$ denote the set of positive integers. Consider a function $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$. For any $m, n \in \mathbb{Z}_{>0}$ we write $f^{n}(m)=\underbrace{f(f(\ldots f}_{n}(m) \ldots))$. Suppose that $f$ has the following two properties:
(i) If $m, n \in \mathbb{Z}_{>0}$, then $\frac{f^{n}(m)-m}{n} \in \mathbb{Z}_{>0}$;
(ii) The set $\mathbb{Z}_{>0} \backslash\left\{f(n) \mid n \in \mathbb{Z}_{>0}\right\}$ is finite.

Prove that the sequence $f(1)-1, f(2)-2, f(3)-3, \ldots$ is periodic.
(Singapore)
N7. Let $\mathbb{Z}_{>0}$ denote the set of positive integers. For any positive integer $k$, a function $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ is called $k$-good if $\operatorname{gcd}(f(m)+n, f(n)+m) \leqslant k$ for all $m \neq n$. Find all $k$ such that there exists a $k$-good function.
(Canada)
N8. For every positive integer $n$ with prime factorization $n=\prod_{i=1}^{k} p_{i}^{\alpha_{i}}$, define

$$
\mho(n)=\sum_{i: p_{i}>10^{100}} \alpha_{i}
$$

That is, $\mho(n)$ is the number of prime factors of $n$ greater than $10^{100}$, counted with multiplicity.
Find all strictly increasing functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that

$$
\mho(f(a)-f(b)) \leqslant \mho(a-b) \quad \text { for all integers } a \text { and } b \text { with } a>b .
$$

## Solutions

## Algebra

A1. Suppose that a sequence $a_{1}, a_{2}, \ldots$ of positive real numbers satisfies

$$
\begin{equation*}
a_{k+1} \geqslant \frac{k a_{k}}{a_{k}^{2}+(k-1)} \tag{1}
\end{equation*}
$$

for every positive integer $k$. Prove that $a_{1}+a_{2}+\cdots+a_{n} \geqslant n$ for every $n \geqslant 2$.
(Serbia)
Solution. From the constraint (1), it can be seen that

$$
\frac{k}{a_{k+1}} \leqslant \frac{a_{k}^{2}+(k-1)}{a_{k}}=a_{k}+\frac{k-1}{a_{k}},
$$

and so

$$
a_{k} \geqslant \frac{k}{a_{k+1}}-\frac{k-1}{a_{k}} .
$$

Summing up the above inequality for $k=1, \ldots, m$, we obtain

$$
\begin{equation*}
a_{1}+a_{2}+\cdots+a_{m} \geqslant\left(\frac{1}{a_{2}}-\frac{0}{a_{1}}\right)+\left(\frac{2}{a_{3}}-\frac{1}{a_{2}}\right)+\cdots+\left(\frac{m}{a_{m+1}}-\frac{m-1}{a_{m}}\right)=\frac{m}{a_{m+1}} \tag{2}
\end{equation*}
$$

Now we prove the problem statement by induction on $n$. The case $n=2$ can be done by applying (1) to $k=1$ :

$$
a_{1}+a_{2} \geqslant a_{1}+\frac{1}{a_{1}} \geqslant 2 .
$$

For the induction step, assume that the statement is true for some $n \geqslant 2$. If $a_{n+1} \geqslant 1$, then the induction hypothesis yields

$$
\begin{equation*}
\left(a_{1}+\cdots+a_{n}\right)+a_{n+1} \geqslant n+1 \tag{3}
\end{equation*}
$$

Otherwise, if $a_{n+1}<1$ then apply (2) as

$$
\left(a_{1}+\cdots+a_{n}\right)+a_{n+1} \geqslant \frac{n}{a_{n+1}}+a_{n+1}=\frac{n-1}{a_{n+1}}+\left(\frac{1}{a_{n+1}}+a_{n+1}\right)>(n-1)+2 .
$$

That completes the solution.
Comment 1. It can be seen easily that having equality in the statement requires $a_{1}=a_{2}=1$ in the base case $n=2$, and $a_{n+1}=1$ in (3). So the equality $a_{1}+\cdots+a_{n}=n$ is possible only in the trivial case $a_{1}=\cdots=a_{n}=1$.

Comment 2. After obtaining (2), there are many ways to complete the solution. We outline three such possibilities.

- With defining $s_{n}=a_{1}+\cdots+a_{n}$, the induction step can be replaced by

$$
s_{n+1}=s_{n}+a_{n+1} \geqslant s_{n}+\frac{n}{s_{n}} \geqslant n+1,
$$

because the function $x \mapsto x+\frac{n}{x}$ increases on $[n, \infty)$.

- By applying the AM-GM inequality to the numbers $a_{1}+\cdots+a_{k}$ and $k a_{k+1}$, we can conclude

$$
a_{1}+\cdots+a_{k}+k a_{k+1} \geqslant 2 k
$$

and sum it up for $k=1, \ldots, n-1$.

- We can derive the symmetric estimate

$$
\sum_{1 \leqslant i<j \leqslant n} a_{i} a_{j}=\sum_{j=2}^{n}\left(a_{1}+\cdots+a_{j-1}\right) a_{j} \geqslant \sum_{j=2}^{n}(j-1)=\frac{n(n-1)}{2}
$$

and combine it with the AM-QM inequality.

A2. Determine all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ with the property that

$$
\begin{equation*}
f(x-f(y))=f(f(x))-f(y)-1 \tag{1}
\end{equation*}
$$

holds for all $x, y \in \mathbb{Z}$.
(Croatia)
Answer. There are two such functions, namely the constant function $x \mapsto-1$ and the successor function $x \mapsto x+1$.

Solution 1. It is immediately checked that both functions mentioned in the answer are as desired.

Now let $f$ denote any function satisfying (1) for all $x, y \in \mathbb{Z}$. Substituting $x=0$ and $y=f(0)$ into (1) we learn that the number $z=-f(f(0))$ satisfies $f(z)=-1$. So by plugging $y=z$ into (1) we deduce that

$$
\begin{equation*}
f(x+1)=f(f(x)) \tag{2}
\end{equation*}
$$

holds for all $x \in \mathbb{Z}$. Thereby (1) simplifies to

$$
\begin{equation*}
f(x-f(y))=f(x+1)-f(y)-1 . \tag{3}
\end{equation*}
$$

We now work towards showing that $f$ is linear by contemplating the difference $f(x+1)-f(x)$ for any $x \in \mathbb{Z}$. By applying (3) with $y=x$ and (2) in this order, we obtain

$$
f(x+1)-f(x)=f(x-f(x))+1=f(f(x-1-f(x)))+1
$$

Since (3) shows $f(x-1-f(x))=f(x)-f(x)-1=-1$, this simplifies to

$$
f(x+1)=f(x)+A
$$

where $A=f(-1)+1$ is some absolute constant.
Now a standard induction in both directions reveals that $f$ is indeed linear and that in fact we have $f(x)=A x+B$ for all $x \in \mathbb{Z}$, where $B=f(0)$. Substituting this into (2) we obtain that

$$
A x+(A+B)=A^{2} x+(A B+B)
$$

holds for all $x \in \mathbb{Z}$; applying this to $x=0$ and $x=1$ we infer $A+B=A B+B$ and $A^{2}=A$. The second equation leads to $A=0$ or $A=1$. In case $A=1$, the first equation gives $B=1$, meaning that $f$ has to be the successor function. If $A=0$, then $f$ is constant and (1) shows that its constant value has to be -1 . Thereby the solution is complete.

Comment. After (2) and (3) have been obtained, there are several other ways to combine them so as to obtain linearity properties of $f$. For instance, using (2) thrice in a row and then (3) with $x=f(y)$ one may deduce that

$$
f(y+2)=f(f(y+1))=f(f(f(y)))=f(f(y)+1)=f(y)+f(0)+1
$$

holds for all $y \in \mathbb{Z}$. It follows that $f$ behaves linearly on the even numbers and on the odd numbers separately, and moreover that the slopes of these two linear functions coincide. From this point, one may complete the solution with some straightforward case analysis.

A different approach using the equations (2) and (3) will be presented in Solution 2. To show that it is also possible to start in a completely different way, we will also present a third solution that avoids these equations entirely.

Solution 2. We commence by deriving (2) and (3) as in the first solution. Now provided that $f$ is injective, (2) tells us that $f$ is the successor function. Thus we may assume from now on that $f$ is not injective, i.e., that there are two integers $a>b$ with $f(a)=f(b)$. A straightforward induction using (2) in the induction step reveals that we have $f(a+n)=f(b+n)$ for all nonnegative integers $n$. Consequently, the sequence $\gamma_{n}=f(b+n)$ is periodic and thus in particular bounded, which means that the numbers

$$
\varphi=\min _{n \geqslant 0} \gamma_{n} \quad \text { and } \quad \psi=\max _{n \geqslant 0} \gamma_{n}
$$

exist.
Let us pick any integer $y$ with $f(y)=\varphi$ and then an integer $x \geqslant a$ with $f(x-f(y))=\varphi$. Due to the definition of $\varphi$ and (3) we have

$$
\varphi \leqslant f(x+1)=f(x-f(y))+f(y)+1=2 \varphi+1
$$

whence $\varphi \geqslant-1$. The same reasoning applied to $\psi$ yields $\psi \leqslant-1$. Since $\varphi \leqslant \psi$ holds trivially, it follows that $\varphi=\psi=-1$, or in other words that we have $f(t)=-1$ for all integers $t \geqslant a$.

Finally, if any integer $y$ is given, we may find an integer $x$ which is so large that $x+1 \geqslant a$ and $x-f(y) \geqslant a$ hold. Due to (3) and the result from the previous paragraph we get

$$
f(y)=f(x+1)-f(x-f(y))-1=(-1)-(-1)-1=-1 .
$$

Thereby the problem is solved.
Solution 3. Set $d=f(0)$. By plugging $x=f(y)$ into (1) we obtain

$$
\begin{equation*}
f^{3}(y)=f(y)+d+1 \tag{4}
\end{equation*}
$$

for all $y \in \mathbb{Z}$, where the left-hand side abbreviates $f(f(f(y)))$. When we replace $x$ in (1) by $f(x)$ we obtain $f(f(x)-f(y))=f^{3}(x)-f(y)-1$ and as a consequence of (4) this simplifies to

$$
\begin{equation*}
f(f(x)-f(y))=f(x)-f(y)+d \tag{5}
\end{equation*}
$$

Now we consider the set

$$
E=\{f(x)-d \mid x \in \mathbb{Z}\} .
$$

Given two integers $a$ and $b$ from $E$, we may pick some integers $x$ and $y$ with $f(x)=a+d$ and $f(y)=b+d$; now (5) tells us that $f(a-b)=(a-b)+d$, which means that $a-b$ itself exemplifies $a-b \in E$. Thus,

$$
\begin{equation*}
E \text { is closed under taking differences. } \tag{6}
\end{equation*}
$$

Also, the definitions of $d$ and $E$ yield $0 \in E$. If $E=\{0\}$, then $f$ is a constant function and (1) implies that the only value attained by $f$ is indeed -1 .

So let us henceforth suppose that $E$ contains some number besides zero. It is known that in this case (6) entails $E$ to be the set of all integer multiples of some positive integer $k$. Indeed, this holds for

$$
k=\min \{|x| \mid x \in E \text { and } x \neq 0\},
$$

as one may verify by an argument based on division with remainder.
Thus we have

$$
\begin{equation*}
\{f(x) \mid x \in \mathbb{Z}\}=\{k \cdot t+d \mid t \in \mathbb{Z}\} \tag{7}
\end{equation*}
$$

Due to (5) and (7) we get

$$
f(k \cdot t)=k \cdot t+d
$$

for all $t \in \mathbb{Z}$, whence in particular $f(k)=k+d$. So by comparing the results of substituting $y=0$ and $y=k$ into (1) we learn that

$$
\begin{equation*}
f(z+k)=f(z)+k \tag{8}
\end{equation*}
$$

holds for all integers $z$. In plain English, this means that on any residue class modulo $k$ the function $f$ is linear with slope 1 .

Now by (7) the set of all values attained by $f$ is such a residue class. Hence, there exists an absolute constant $c$ such that $f(f(x))=f(x)+c$ holds for all $x \in \mathbb{Z}$. Thereby (1) simplifies to

$$
\begin{equation*}
f(x-f(y))=f(x)-f(y)+c-1 \tag{9}
\end{equation*}
$$

On the other hand, considering (1) modulo $k$ we obtain $d \equiv-1(\bmod k)$ because of (7). So by (7) again, $f$ attains the value -1 .

Thus we may apply (9) to some integer $y$ with $f(y)=-1$, which gives $f(x+1)=f(x)+c$. So $f$ is a linear function with slope $c$. Hence, (8) leads to $c=1$, wherefore there is an absolute constant $d^{\prime}$ with $f(x)=x+d^{\prime}$ for all $x \in \mathbb{Z}$. Using this for $x=0$ we obtain $d^{\prime}=d$ and finally (4) discloses $d=1$, meaning that $f$ is indeed the successor function.

A3. Let $n$ be a fixed positive integer. Find the maximum possible value of

$$
\sum_{1 \leqslant r<s \leqslant 2 n}(s-r-n) x_{r} x_{s},
$$

where $-1 \leqslant x_{i} \leqslant 1$ for all $i=1,2, \ldots, 2 n$.

Answer. $n(n-1)$.
Solution 1. Let $Z$ be the expression to be maximized. Since this expression is linear in every variable $x_{i}$ and $-1 \leqslant x_{i} \leqslant 1$, the maximum of $Z$ will be achieved when $x_{i}=-1$ or 1 . Therefore, it suffices to consider only the case when $x_{i} \in\{-1,1\}$ for all $i=1,2, \ldots, 2 n$.

For $i=1,2, \ldots, 2 n$, we introduce auxiliary variables

$$
y_{i}=\sum_{r=1}^{i} x_{r}-\sum_{r=i+1}^{2 n} x_{r} .
$$

Taking squares of both sides, we have

$$
\begin{align*}
y_{i}^{2} & =\sum_{r=1}^{2 n} x_{r}^{2}+\sum_{r<s \leqslant i} 2 x_{r} x_{s}+\sum_{i<r<s} 2 x_{r} x_{s}-\sum_{r \leqslant i<s} 2 x_{r} x_{s} \\
& =2 n+\sum_{r<s \leqslant i} 2 x_{r} x_{s}+\sum_{i<r<s} 2 x_{r} x_{s}-\sum_{r \leqslant i<s} 2 x_{r} x_{s}, \tag{1}
\end{align*}
$$

where the last equality follows from the fact that $x_{r} \in\{-1,1\}$. Notice that for every $r<s$, the coefficient of $x_{r} x_{s}$ in (1) is 2 for each $i=1, \ldots, r-1, s, \ldots, 2 n$, and this coefficient is -2 for each $i=r, \ldots, s-1$. This implies that the coefficient of $x_{r} x_{s}$ in $\sum_{i=1}^{2 n} y_{i}^{2}$ is $2(2 n-s+r)-2(s-r)=$ $4(n-s+r)$. Therefore, summing (1) for $i=1,2, \ldots, 2 n$ yields

$$
\begin{equation*}
\sum_{i=1}^{2 n} y_{i}^{2}=4 n^{2}+\sum_{1 \leqslant r<s \leqslant 2 n} 4(n-s+r) x_{r} x_{s}=4 n^{2}-4 Z \tag{2}
\end{equation*}
$$

Hence, it suffices to find the minimum of the left-hand side.
Since $x_{r} \in\{-1,1\}$, we see that $y_{i}$ is an even integer. In addition, $y_{i}-y_{i-1}=2 x_{i}= \pm 2$, and so $y_{i-1}$ and $y_{i}$ are consecutive even integers for every $i=2,3, \ldots, 2 n$. It follows that $y_{i-1}^{2}+y_{i}^{2} \geqslant 4$, which implies

$$
\begin{equation*}
\sum_{i=1}^{2 n} y_{i}^{2}=\sum_{j=1}^{n}\left(y_{2 j-1}^{2}+y_{2 j}^{2}\right) \geqslant 4 n \tag{3}
\end{equation*}
$$

Combining (2) and (3), we get

$$
\begin{equation*}
4 n \leqslant \sum_{i=1}^{2 n} y_{i}^{2}=4 n^{2}-4 Z \tag{4}
\end{equation*}
$$

Hence, $Z \leqslant n(n-1)$.
If we set $x_{i}=1$ for odd indices $i$ and $x_{i}=-1$ for even indices $i$, then we obtain equality in (3) (and thus in (4)). Therefore, the maximum possible value of $Z$ is $n(n-1)$, as desired.

Comment 1. $Z=n(n-1)$ can be achieved by several other examples. In particular, $x_{i}$ needs not be $\pm 1$. For instance, setting $x_{i}=(-1)^{i}$ for all $2 \leqslant i \leqslant 2 n$, we find that the coefficient of $x_{1}$ in $Z$ is 0 . Therefore, $x_{1}$ can be chosen arbitrarily in the interval $[-1,1]$.

Nevertheless, if $x_{i} \in\{-1,1\}$ for all $i=1,2, \ldots, 2 n$, then the equality $Z=n(n-1)$ holds only when $\left(y_{1}, y_{2}, \ldots, y_{2 n}\right)=(0, \pm 2,0, \pm 2, \ldots, 0, \pm 2)$ or $( \pm 2,0, \pm 2,0, \ldots, \pm 2,0)$. In each case, we can reconstruct $x_{i}$ accordingly. The sum $\sum_{i=1}^{2 n} x_{i}$ in the optimal cases needs not be 0 , but it must equal 0 or $\pm 2$.

Comment 2. Several variations in setting up the auxiliary variables are possible. For instance, one may let $x_{2 n+i}=-x_{i}$ and $y_{i}^{\prime}=x_{i}+x_{i+1}+\cdots+x_{i+n-1}$ for any $1 \leqslant i \leqslant 2 n$. Similarly to Solution 1 , we obtain $Y:=y_{1}^{\prime 2}+y_{2}^{\prime 2}+\cdots+y_{2 n}^{\prime 2}=2 n^{2}-2 Z$. Then, it suffices to show that $Y \geqslant 2 n$. If $n$ is odd, then each $y_{i}^{\prime}$ is odd, and so $y_{i}^{\prime 2} \geqslant 1$. If $n$ is even, then each $y_{i}^{\prime}$ is even. We can check that at least one of $y_{i}^{\prime}, y_{i+1}^{\prime}, y_{n+i}^{\prime}$, and $y_{n+i+1}^{\prime}$ is nonzero, so that $y_{i}^{\prime 2}+y_{i+1}^{\prime 2}+y_{n+i}^{\prime 2}+y_{n+i+1}^{\prime 2} \geqslant 4$; summing these up for $i=1,3, \ldots, n-1$ yields $Y \geqslant 2 n$.

Solution 2. We present a different method of obtaining the bound $Z \leqslant n(n-1)$. As in the previous solution, we reduce the problem to the case $x_{i} \in\{-1,1\}$. For brevity, we use the notation $[2 n]=\{1,2, \ldots, 2 n\}$.

Consider any $x_{1}, x_{2}, \ldots, x_{2 n} \in\{-1,1\}$. Let

$$
A=\left\{i \in[2 n]: x_{i}=1\right\} \quad \text { and } \quad B=\left\{i \in[2 n]: x_{i}=-1\right\}
$$

For any subsets $X$ and $Y$ of [2n] we define

$$
e(X, Y)=\sum_{r<s, r \in X, s \in Y}(s-r-n)
$$

One may observe that
$e(A, A)+e(A, B)+e(B, A)+e(B, B)=e([2 n],[2 n])=\sum_{1 \leqslant r<s \leqslant 2 n}(s-r-n)=-\frac{(n-1) n(2 n-1)}{3}$.
Therefore, we have

$$
\begin{equation*}
Z=e(A, A)-e(A, B)-e(B, A)+e(B, B)=2(e(A, A)+e(B, B))+\frac{(n-1) n(2 n-1)}{3} \tag{5}
\end{equation*}
$$

Thus, we need to maximize $e(A, A)+e(B, B)$, where $A$ and $B$ form a partition of [2n].
Due to the symmetry, we may assume that $|A|=n-p$ and $|B|=n+p$, where $0 \leqslant p \leqslant n$. From now on, we fix the value of $p$ and find an upper bound for $Z$ in terms of $n$ and $p$.

Let $a_{1}<a_{2}<\cdots<a_{n-p}$ and $b_{1}<b_{2}<\cdots<b_{n+p}$ list all elements of $A$ and $B$, respectively. Then

$$
\begin{equation*}
e(A, A)=\sum_{1 \leqslant i<j \leqslant n-p}\left(a_{j}-a_{i}-n\right)=\sum_{i=1}^{n-p}(2 i-1-n+p) a_{i}-\binom{n-p}{2} \cdot n \tag{6}
\end{equation*}
$$

and similarly

$$
\begin{equation*}
e(B, B)=\sum_{i=1}^{n+p}(2 i-1-n-p) b_{i}-\binom{n+p}{2} \cdot n \tag{7}
\end{equation*}
$$

Thus, now it suffices to maximize the value of

$$
\begin{equation*}
M=\sum_{i=1}^{n-p}(2 i-1-n+p) a_{i}+\sum_{i=1}^{n+p}(2 i-1-n-p) b_{i} \tag{8}
\end{equation*}
$$

In order to get an upper bound, we will apply the rearrangement inequality to the sequence $a_{1}, a_{2}, \ldots, a_{n-p}, b_{1}, b_{2}, \ldots, b_{n+p}$ (which is a permutation of $1,2, \ldots, 2 n$ ), together with the sequence of coefficients of these numbers in (8). The coefficients of $a_{i}$ form the sequence

$$
n-p-1, n-p-3, \ldots, 1-n+p
$$

and those of $b_{i}$ form the sequence

$$
n+p-1, n+p-3, \ldots, 1-n-p .
$$

Altogether, these coefficients are, in descending order:

- $n+p+1-2 i$, for $i=1,2, \ldots, p$;
- $n-p+1-2 i$, counted twice, for $i=1,2, \ldots, n-p$; and
- $-(n+p+1-2 i)$, for $i=p, p-1, \ldots, 1$.

Thus, the rearrangement inequality yields

$$
\begin{align*}
& M \leqslant \sum_{i=1}^{p}(n+p+1-2 i)(2 n+1-i) \\
& \quad+\sum_{i=1}^{n-p}(n-p+1-2 i)((2 n+2-p-2 i)+(2 n+1-p-2 i)) \\
& \quad-\sum_{i=1}^{p}(n+p+1-2 i) i . \tag{9}
\end{align*}
$$

Finally, combining the information from (5), (6), (7), and (9), we obtain

$$
\begin{aligned}
Z \leqslant & \frac{(n-1) n(2 n-1)}{3}-2 n\left(\binom{n-p}{2}+\binom{n+p}{2}\right) \\
& +2 \sum_{i=1}^{p}(n+p+1-2 i)(2 n+1-2 i)+2 \sum_{i=1}^{n-p}(n-p+1-2 i)(4 n-2 p+3-4 i),
\end{aligned}
$$

which can be simplified to

$$
Z \leqslant n(n-1)-\frac{2}{3} p(p-1)(p+1)
$$

Since $p$ is a nonnegative integer, this yields $Z \leqslant n(n-1)$.

A4. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ satisfying the equation

$$
\begin{equation*}
f(x+f(x+y))+f(x y)=x+f(x+y)+y f(x) \tag{1}
\end{equation*}
$$

for all real numbers $x$ and $y$.
(Albania)
Answer. There are two such functions, namely the identity function and $x \mapsto 2-x$.
Solution. Clearly, each of the functions $x \mapsto x$ and $x \mapsto 2-x$ satisfies (1). It suffices now to show that they are the only solutions to the problem.

Suppose that $f$ is any function satisfying (1). Then setting $y=1$ in (1), we obtain

$$
\begin{equation*}
f(x+f(x+1))=x+f(x+1) \tag{2}
\end{equation*}
$$

in other words, $x+f(x+1)$ is a fixed point of $f$ for every $x \in \mathbb{R}$.
We distinguish two cases regarding the value of $f(0)$.
Case 1. $\quad f(0) \neq 0$.
By letting $x=0$ in (1), we have

$$
f(f(y))+f(0)=f(y)+y f(0) .
$$

So, if $y_{0}$ is a fixed point of $f$, then substituting $y=y_{0}$ in the above equation we get $y_{0}=1$. Thus, it follows from (2) that $x+f(x+1)=1$ for all $x \in \mathbb{R}$. That is, $f(x)=2-x$ for all $x \in \mathbb{R}$. Case 2. $\quad f(0)=0$.

By letting $y=0$ and replacing $x$ by $x+1$ in (1), we obtain

$$
\begin{equation*}
f(x+f(x+1)+1)=x+f(x+1)+1 \tag{3}
\end{equation*}
$$

From (1), the substitution $x=1$ yields

$$
\begin{equation*}
f(1+f(y+1))+f(y)=1+f(y+1)+y f(1) . \tag{4}
\end{equation*}
$$

By plugging $x=-1$ into (2), we see that $f(-1)=-1$. We then plug $y=-1$ into (4) and deduce that $f(1)=1$. Hence, (4) reduces to

$$
\begin{equation*}
f(1+f(y+1))+f(y)=1+f(y+1)+y \tag{5}
\end{equation*}
$$

Accordingly, if both $y_{0}$ and $y_{0}+1$ are fixed points of $f$, then so is $y_{0}+2$. Thus, it follows from (2) and (3) that $x+f(x+1)+2$ is a fixed point of $f$ for every $x \in \mathbb{R}$; i.e.,

$$
f(x+f(x+1)+2)=x+f(x+1)+2
$$

Replacing $x$ by $x-2$ simplifies the above equation to

$$
f(x+f(x-1))=x+f(x-1)
$$

On the other hand, we set $y=-1$ in (1) and get

$$
f(x+f(x-1))=x+f(x-1)-f(x)-f(-x)
$$

Therefore, $f(-x)=-f(x)$ for all $x \in \mathbb{R}$.
Finally, we substitute $(x, y)$ by $(-1,-y)$ in (1) and use the fact that $f(-1)=-1$ to get

$$
f(-1+f(-y-1))+f(y)=-1+f(-y-1)+y
$$

Since $f$ is an odd function, the above equation becomes

$$
-f(1+f(y+1))+f(y)=-1-f(y+1)+y
$$

By adding this equation to (5), we conclude that $f(y)=y$ for all $y \in \mathbb{R}$.

A5. Let $2 \mathbb{Z}+1$ denote the set of odd integers. Find all functions $f: \mathbb{Z} \rightarrow 2 \mathbb{Z}+1$ satisfying

$$
\begin{equation*}
f(x+f(x)+y)+f(x-f(x)-y)=f(x+y)+f(x-y) \tag{1}
\end{equation*}
$$

for every $x, y \in \mathbb{Z}$.
(U.S.A.)

Answer. Fix an odd positive integer $d$, an integer $k$, and odd integers $\ell_{0}, \ell_{1}, \ldots, \ell_{d-1}$. Then the function defined as

$$
f(m d+i)=2 k m d+\ell_{i} d \quad(m \in \mathbb{Z}, \quad i=0,1, \ldots, d-1)
$$

satisfies the problem requirements, and these are all such functions.
Solution. Throughout the solution, all functions are assumed to map integers to integers.
For any function $g$ and any nonzero integer $t$, define

$$
\Delta_{t} g(x)=g(x+t)-g(x)
$$

For any nonzero integers $a$ and $b$, notice that $\Delta_{a} \Delta_{b} g=\Delta_{b} \Delta_{a} g$. Moreover, if $\Delta_{a} g=0$ and $\Delta_{b} g=0$, then $\Delta_{a+b} g=0$ and $\Delta_{a t} g=0$ for all nonzero integers $t$. We say that $g$ is $t$-quasiperiodic if $\Delta_{t} g$ is a constant function (in other words, if $\Delta_{1} \Delta_{t} g=0$, or $\Delta_{1} g$ is $t$-periodic). In this case, we call $t$ a quasi-period of $g$. We say that $g$ is quasi-periodic if it is $t$-quasi-periodic for some nonzero integer $t$.

Notice that a quasi-period of $g$ is a period of $\Delta_{1} g$. So if $g$ is quasi-periodic, then its minimal positive quasi-period $t$ divides all its quasi-periods.

We now assume that $f$ satisfies (1). First, by setting $a=x+y$, the problem condition can be rewritten as

$$
\begin{equation*}
\Delta_{f(x)} f(a)=\Delta_{f(x)} f(2 x-a-f(x)) \quad \text { for all } x, a \in \mathbb{Z} \tag{2}
\end{equation*}
$$

Let $b$ be an arbitrary integer and let $k$ be an arbitrary positive integer. Applying (2) when $a$ is substituted by $b, b+f(x), \ldots, b+(k-1) f(x)$ and summing up all these equations, we get

$$
\Delta_{k f(x)} f(b)=\Delta_{k f(x)} f(2 x-b-k f(x)) .
$$

Notice that a similar argument works when $k$ is negative, so that

$$
\begin{equation*}
\Delta_{M} f(b)=\Delta_{M} f(2 x-b-M) \quad \text { for any nonzero integer } M \text { such that } f(x) \mid M \tag{3}
\end{equation*}
$$

We now prove two lemmas.
Lemma 1. For any distinct integers $x$ and $y$, the function $\Delta_{\operatorname{lcm}(f(x), f(y))} f$ is $2(y-x)$-periodic. Proof. Denote $L=\operatorname{lcm}(f(x), f(y))$. Applying (3) twice, we obtain

$$
\Delta_{L} f(b)=\Delta_{L} f(2 x-b-L)=\Delta_{L} f(2 y-(b+2(y-x))-L)=\Delta_{L} f(b+2(y-x))
$$

Thus, the function $\Delta_{L} f$ is $2(y-x)$-periodic, as required.
Lemma 2. Let $g$ be a function. If $t$ and $s$ are nonzero integers such that $\Delta_{t s} g=0$ and $\Delta_{t} \Delta_{t} g=0$, then $\Delta_{t} g=0$.
Proof. Assume, without loss of generality, that $s$ is positive. Let $a$ be an arbitrary integer. Since $\Delta_{t} \Delta_{t} g=0$, we have

$$
\Delta_{t} g(a)=\Delta_{t} g(a+t)=\cdots=\Delta_{t} g(a+(s-1) t)
$$

The sum of these $s$ equal numbers is $\Delta_{t s} g(a)=0$, so each of them is zero, as required.

We now return to the solution.
Step 1. We prove that $f$ is quasi-periodic.
Let $Q=\operatorname{lcm}(f(0), f(1))$. Applying Lemma 1, we get that the function $g=\Delta_{Q} f$ is 2-periodic. In other words, the values of $g$ are constant on even numbers and on odd numbers separately. Moreover, setting $M=Q$ and $x=b=0$ in (3), we get $g(0)=g(-Q)$. Since 0 and $-Q$ have different parities, the value of $g$ at even numbers is the same as that at odd numbers. Thus, $g$ is constant, which means that $Q$ is a quasi-period of $f$.
Step 2. Denote the minimal positive quasi-period of $f$ by $T$. We prove that $T \mid f(x)$ for all integers $x$.

Since an odd number $Q$ is a quasi-period of $f$, the number $T$ is also odd. Now suppose, to the contrary, that there exist an odd prime $p$, a positive integer $\alpha$, and an integer $u$ such that $p^{\alpha} \mid T$ but $p^{\alpha} \nmid f(u)$. Setting $x=u$ and $y=0$ in (1), we have $2 f(u)=f(u+f(u))+f(u-f(u))$, so $p^{\alpha}$ does not divide the value of $f$ at one of the points $u+f(u)$ or $u-f(u)$. Denote this point by $v$.

Let $L=\operatorname{lcm}(f(u), f(v))$. Since $|u-v|=f(u)$, from Lemma 1 we get $\Delta_{2 f(u)} \Delta_{L} f=0$. Hence the function $\Delta_{L} f$ is $2 f(u)$-periodic as well as $T$-periodic, so it is $\operatorname{gcd}(T, 2 f(u))$-periodic, or $\Delta_{\operatorname{gcd}(T, 2 f(u))} \Delta_{L} f=0$. Similarly, observe that the function $\Delta_{\operatorname{gcd}(T, 2 f(u))} f$ is $L$-periodic as well as $T$-periodic, so we may conclude that $\Delta_{\operatorname{gcd}(T, L)} \Delta_{\operatorname{gcd}(T, 2 f(u))} f=0$. Since $p^{\alpha} \nmid L$, both $\operatorname{gcd}(T, 2 f(u))$ and $\operatorname{gcd}(T, L)$ divide $T / p$. We thus obtain $\Delta_{T / p} \Delta_{T / p} f=0$, which yields

$$
\Delta_{T / p} \Delta_{T / p} \Delta_{1} f=0
$$

Since $\Delta_{T} \Delta_{1} f=0$, we can apply Lemma 2 to the function $\Delta_{1} f$, obtaining $\Delta_{T / p} \Delta_{1} f=0$. However, this means that $f$ is $(T / p)$-quasi-periodic, contradicting the minimality of $T$. Our claim is proved.

## Step 3. We describe all functions $f$.

Let $d$ be the greatest common divisor of all values of $f$. Then $d$ is odd. By Step $2, d$ is a quasi-period of $f$, so that $\Delta_{d} f$ is constant. Since the value of $\Delta_{d} f$ is even and divisible by $d$, we may denote this constant by $2 d k$, where $k$ is an integer. Next, for all $i=0,1, \ldots, d-1$, define $\ell_{i}=f(i) / d$; notice that $\ell_{i}$ is odd. Then

$$
f(m d+i)=\Delta_{m d} f(i)+f(i)=2 k m d+\ell_{i} d \quad \text { for all } m \in \mathbb{Z} \quad \text { and } i=0,1, \ldots, d-1
$$

This shows that all functions satisfying (1) are listed in the answer.
It remains to check that all such functions indeed satisfy (1). This is equivalent to checking (2), which is true because for every integer $x$, the value of $f(x)$ is divisible by $d$, so that $\Delta_{f(x)} f$ is constant.

Comment. After obtaining Lemmas 1 and 2, it is possible to complete the steps in a different order. Here we sketch an alternative approach.

For any function $g$ and any nonzero integer $t$, we say that $g$ is $t$-pseudo-periodic if $\Delta_{t} \Delta_{t} g=0$. In this case, we call $t$ a pseudo-period of $g$, and we say that $g$ is pseudo-periodic.

Let us first prove a basic property: if a function $g$ is pseudo-periodic, then its minimal positive pseudo-period divides all its pseudo-periods. To establish this, it suffices to show that if $t$ and $s$ are pseudo-periods of $g$ with $t \neq s$, then so is $t-s$. Indeed, suppose that $\Delta_{t} \Delta_{t} g=\Delta_{s} \Delta_{s} g=0$. Then $\Delta_{t} \Delta_{t} \Delta_{s} g=\Delta_{t s} \Delta_{s} g=0$, so that $\Delta_{t} \Delta_{s} g=0$ by Lemma 2. Taking differences, we obtain $\Delta_{t} \Delta_{t-s} g=\Delta_{s} \Delta_{t-s} g=0$, and thus $\Delta_{t-s} \Delta_{t-s} g=0$.

Now let $f$ satisfy the problem condition. We will show that $f$ is pseudo-periodic. When this is done, we will let $T^{\prime}$ be the minimal pseudo-period of $f$, and show that $T^{\prime}$ divides $2 f(x)$ for every integer $x$, using arguments similar to Step 2 of the solution. Then we will come back to Step 1 by showing that $T^{\prime}$ is also a quasi-period of $f$.

First, Lemma 1 yields that $\Delta_{2(y-x)} \Delta_{\operatorname{lcm}(f(x), f(y))} f=0$ for every distinct integers $x$ and $y$. Hence $f$ is pseudo-periodic with pseudo-period $L_{x, y}=\operatorname{lcm}(2(y-x), f(x), f(y))$.

We now show that $T^{\prime} \mid 2 f(x)$ for every integer $x$. Suppose, to the contrary, that there exists an integer $u$, a prime $p$, and a positive integer $\alpha$ such that $p^{\alpha} \mid T^{\prime}$ and $p^{\alpha} \nmid 2 f(u)$. Choose $v$ as in Step 2 and employ Lemma 1 to obtain $\Delta_{2 f(u)} \Delta_{\operatorname{lcm}(f(u), f(v))} f=0$. However, this implies that $\Delta_{T^{\prime} / p} \Delta_{T^{\prime} / p} f=0$, a contradiction with the minimality of $T^{\prime}$.

We now claim that $\Delta_{T^{\prime}} \Delta_{2} f=0$. Indeed, Lemma 1 implies that there exists an integer $s$ such that $\Delta_{s} \Delta_{2} f=0$. Hence $\Delta_{T^{\prime} s} \Delta_{2} f=\Delta_{T^{\prime}} \Delta_{T^{\prime}} \Delta_{2} f=0$, which allows us to conclude that $\Delta_{T^{\prime}} \Delta_{2} f=0$ by Lemma 2. (The last two paragraphs are similar to Step 2 of the solution.)

Now, it is not difficult to finish the solution, though more work is needed to eliminate the factors of 2 from the subscripts of $\Delta_{T^{\prime}} \Delta_{2} f=0$. Once this is done, we will obtain an odd quasi-period of $f$ that divides $f(x)$ for all integers $x$. Then we can complete the solution as in Step 3.

A6. Let $n$ be a fixed integer with $n \geqslant 2$. We say that two polynomials $P$ and $Q$ with real coefficients are block-similar if for each $i \in\{1,2, \ldots, n\}$ the sequences

$$
\begin{aligned}
& P(2015 i), P(2015 i-1), \ldots, P(2015 i-2014) \quad \text { and } \\
& Q(2015 i), Q(2015 i-1), \ldots, Q(2015 i-2014)
\end{aligned}
$$

are permutations of each other.
(a) Prove that there exist distinct block-similar polynomials of degree $n+1$.
(b) Prove that there do not exist distinct block-similar polynomials of degree $n$.
(Canada)
Solution 1. For convenience, we set $k=2015=2 \ell+1$.
Part (a). Consider the following polynomials of degree $n+1$ :

$$
P(x)=\prod_{i=0}^{n}(x-i k) \quad \text { and } \quad Q(x)=\prod_{i=0}^{n}(x-i k-1) .
$$

Since $Q(x)=P(x-1)$ and $P(0)=P(k)=P(2 k)=\cdots=P(n k)$, these polynomials are block-similar (and distinct).

Part (b). For every polynomial $F(x)$ and every nonnegative integer $m$, define $\Sigma_{F}(m)=$ $\sum_{i=1}^{m} F(i)$; in particular, $\Sigma_{F}(0)=0$. It is well-known that for every nonnegative integer $d$ the sum $\sum_{i=1}^{m} i^{d}$ is a polynomial in $m$ of degree $d+1$. Thus $\Sigma_{F}$ may also be regarded as a real polynomial of degree $\operatorname{deg} F+1$ (with the exception that if $F=0$, then $\Sigma_{F}=0$ as well). This allows us to consider the values of $\Sigma_{F}$ at all real points (where the initial definition does not apply).

Assume for the sake of contradiction that there exist two distinct block-similar polynomials $P(x)$ and $Q(x)$ of degree $n$. Then both polynomials $\Sigma_{P-Q}(x)$ and $\Sigma_{P^{2}-Q^{2}}(x)$ have roots at the points $0, k, 2 k, \ldots, n k$. This motivates the following lemma, where we use the special polynomial

$$
T(x)=\prod_{i=0}^{n}(x-i k)
$$

Lemma. Assume that $F(x)$ is a nonzero polynomial such that $0, k, 2 k, \ldots, n k$ are among the roots of the polynomial $\Sigma_{F}(x)$. Then $\operatorname{deg} F \geqslant n$, and there exists a polynomial $G(x)$ such that $\operatorname{deg} G=\operatorname{deg} F-n$ and $F(x)=T(x) G(x)-T(x-1) G(x-1)$.
Proof. If $\operatorname{deg} F<n$, then $\Sigma_{F}(x)$ has at least $n+1$ roots, while its degree is less than $n+1$. Therefore, $\Sigma_{F}(x)=0$ and hence $F(x)=0$, which is impossible. Thus $\operatorname{deg} F \geqslant n$.

The lemma condition yields that $\Sigma_{F}(x)=T(x) G(x)$ for some polynomial $G(x)$ such that $\operatorname{deg} G=\operatorname{deg} \Sigma_{F}-(n+1)=\operatorname{deg} F-n$.

Now, let us define $F_{1}(x)=T(x) G(x)-T(x-1) G(x-1)$. Then for every positive integer $n$ we have

$$
\Sigma_{F_{1}}(n)=\sum_{i=1}^{n}(T(x) G(x)-T(x-1) G(x-1))=T(n) G(n)-T(0) G(0)=T(n) G(n)=\Sigma_{F}(n)
$$

so the polynomial $\Sigma_{F-F_{1}}(x)=\Sigma_{F}(x)-\Sigma_{F_{1}}(x)$ has infinitely many roots. This means that this polynomial is zero, which in turn yields $F(x)=F_{1}(x)$, as required.

First, we apply the lemma to the nonzero polynomial $R_{1}(x)=P(x)-Q(x)$. Since the degree of $R_{1}(x)$ is at most $n$, we conclude that it is exactly $n$. Moreover, $R_{1}(x)=\alpha \cdot(T(x)-T(x-1))$ for some nonzero constant $\alpha$.

Our next aim is to prove that the polynomial $S(x)=P(x)+Q(x)$ is constant. Assume the contrary. Then, notice that the polynomial $R_{2}(x)=P(x)^{2}-Q(x)^{2}=R_{1}(x) S(x)$ is also nonzero and satisfies the lemma condition. Since $n<\operatorname{deg} R_{1}+\operatorname{deg} S=\operatorname{deg} R_{2} \leqslant 2 n$, the lemma yields

$$
R_{2}(x)=T(x) G(x)-T(x-1) G(x-1)
$$

with some polynomial $G(x)$ with $0<\operatorname{deg} G \leqslant n$.
Since the polynomial $R_{1}(x)=\alpha(T(x)-T(x-1))$ divides the polynomial

$$
R_{2}(x)=T(x)(G(x)-G(x-1))+G(x-1)(T(x)-T(x-1)),
$$

we get $R_{1}(x) \mid T(x)(G(x)-G(x-1))$. On the other hand,

$$
\operatorname{gcd}\left(T(x), R_{1}(x)\right)=\operatorname{gcd}(T(x), T(x)-T(x-1))=\operatorname{gcd}(T(x), T(x-1))=1
$$

since both $T(x)$ and $T(x-1)$ are the products of linear polynomials, and their roots are distinct. Thus $R_{1}(x) \mid G(x)-G(x-1)$. However, this is impossible since $G(x)-G(x-1)$ is a nonzero polynomial of degree less than $n=\operatorname{deg} R_{1}$.

Thus, our assumption is wrong, and $S(x)$ is a constant polynomial, say $S(x)=\beta$. Notice that the polynomials $(2 P(x)-\beta) / \alpha$ and $(2 Q(x)-\beta) / \alpha$ are also block-similar and distinct. So we may replace the initial polynomials by these ones, thus obtaining two block-similar polynomials $P(x)$ and $Q(x)$ with $P(x)=-Q(x)=T(x)-T(x-1)$. It remains to show that this is impossible.

For every $i=1,2 \ldots, n$, the values $T(i k-k+1)$ and $T(i k-1)$ have the same sign. This means that the values $P(i k-k+1)=T(i k-k+1)$ and $P(i k)=-T(i k-1)$ have opposite signs, so $P(x)$ has a root in each of the $n$ segments $[i k-k+1, i k]$. Since $\operatorname{deg} P=n$, it must have exactly one root in each of them.

Thus, the sequence $P(1), P(2), \ldots, P(k)$ should change sign exactly once. On the other hand, since $P(x)$ and $-P(x)$ are block-similar, this sequence must have as many positive terms as negative ones. Since $k=2 \ell+1$ is odd, this shows that the middle term of the sequence above must be zero, so $P(\ell+1)=0$, or $T(\ell+1)=T(\ell)$. However, this is not true since

$$
|T(\ell+1)|=|\ell+1| \cdot|\ell| \cdot \prod_{i=2}^{n}|\ell+1-i k|<|\ell| \cdot|\ell+1| \cdot \prod_{i=2}^{n}|\ell-i k|=|T(\ell)|
$$

where the strict inequality holds because $n \geqslant 2$. We come to the final contradiction.

Comment 1. In the solution above, we used the fact that $k>1$ is odd. One can modify the arguments of the last part in order to work for every (not necessarily odd) sufficiently large value of $k$; namely, when $k$ is even, one may show that the sequence $P(1), P(2), \ldots, P(k)$ has different numbers of positive and negative terms.

On the other hand, the problem statement with $k$ replaced by 2 is false, since the polynomials $P(x)=T(x)-T(x-1)$ and $Q(x)=T(x-1)-T(x)$ are block-similar in this case, due to the fact that $P(2 i-1)=-P(2 i)=Q(2 i)=-Q(2 i-1)=T(2 i-1)$ for all $i=1,2, \ldots, n$. Thus, every complete solution should use the relation $k>2$.

One may easily see that the condition $n \geqslant 2$ is also substantial, since the polynomials $x$ and $k+1-x$ become block-similar if we set $n=1$.

It is easily seen from the solution that the result still holds if we assume that the polynomials have degree at most $n$.

Solution 2. We provide an alternative argument for part (b).
Assume again that there exist two distinct block-similar polynomials $P(x)$ and $Q(x)$ of degree $n$. Let $R(x)=P(x)-Q(x)$ and $S(x)=P(x)+Q(x)$. For brevity, we also denote the segment $[(i-1) k+1, i k]$ by $I_{i}$, and the set $\{(i-1) k+1,(i-1) k+2, \ldots, i k\}$ of all integer points in $I_{i}$ by $Z_{i}$.
Step 1. We prove that $R(x)$ has exactly one root in each segment $I_{i}, i=1,2, \ldots, n$, and all these roots are simple.

Indeed, take any $i \in\{1,2, \ldots, n\}$ and choose some points $p^{-}, p^{+} \in Z_{i}$ so that

$$
P\left(p^{-}\right)=\min _{x \in Z_{i}} P(x) \quad \text { and } \quad P\left(p^{+}\right)=\max _{x \in Z_{i}} P(x) .
$$

Since the sequences of values of $P$ and $Q$ in $Z_{i}$ are permutations of each other, we have $R\left(p^{-}\right)=P\left(p^{-}\right)-Q\left(p^{-}\right) \leqslant 0$ and $R\left(p^{+}\right)=P\left(p^{+}\right)-Q\left(p^{+}\right) \geqslant 0$. Since $R(x)$ is continuous, there exists at least one root of $R(x)$ between $p^{-}$and $p^{+}$- thus in $I_{i}$.

So, $R(x)$ has at least one root in each of the $n$ disjoint segments $I_{i}$ with $i=1,2, \ldots, n$. Since $R(x)$ is nonzero and its degree does not exceed $n$, it should have exactly one root in each of these segments, and all these roots are simple, as required.

Step 2. We prove that $S(x)$ is constant.
We start with the following claim.
Claim. For every $i=1,2, \ldots, n$, the sequence of values $S((i-1) k+1), S((i-1) k+2), \ldots$, $S(i k)$ cannot be strictly increasing.
Proof. Fix any $i \in\{1,2, \ldots, n\}$. Due to the symmetry, we may assume that $P(i k) \leqslant Q(i k)$. Choose now $p^{-}$and $p^{+}$as in Step 1. If we had $P\left(p^{+}\right)=P\left(p^{-}\right)$, then $P$ would be constant on $Z_{i}$, so all the elements of $Z_{i}$ would be the roots of $R(x)$, which is not the case. In particular, we have $p^{+} \neq p^{-}$. If $p^{-}>p^{+}$, then $S\left(p^{-}\right)=P\left(p^{-}\right)+Q\left(p^{-}\right) \leqslant Q\left(p^{+}\right)+P\left(p^{+}\right)=S\left(p^{+}\right)$, so our claim holds.

We now show that the remaining case $p^{-}<p^{+}$is impossible. Assume first that $P\left(p^{+}\right)>$ $Q\left(p^{+}\right)$. Then, like in Step 1, we have $R\left(p^{-}\right) \leqslant 0, R\left(p^{+}\right)>0$, and $R(i k) \leqslant 0$, so $R(x)$ has a root in each of the intervals $\left[p^{-}, p^{+}\right)$and $\left(p^{+}, i k\right]$. This contradicts the result of Step 1.

We are left only with the case $p^{-}<p^{+}$and $P\left(p^{+}\right)=Q\left(p^{+}\right)$(thus $p^{+}$is the unique root of $R(x)$ in $\left.I_{i}\right)$. If $p^{+}=i k$, then the values of $R(x)$ on $Z_{i} \backslash\{i k\}$ are all of the same sign, which is absurd since their sum is zero. Finally, if $p^{-}<p^{+}<i k$, then $R\left(p^{-}\right)$and $R(i k)$ are both negative. This means that $R(x)$ should have an even number of roots in $\left[p^{-}, i k\right]$, counted with multiplicity. This also contradicts the result of Step 1.

In a similar way, one may prove that for every $i=1,2, \ldots, n$, the sequence $S((i-1) k+1)$, $S((i-1) k+2), \ldots, S(i k)$ cannot be strictly decreasing. This means that the polynomial $\Delta S(x)=S(x)-S(x-1)$ attains at least one nonnegative value, as well as at least one nonpositive value, on the set $Z_{i}$ (and even on $Z_{i} \backslash\{(i-1) k+1\}$ ); so $\Delta S$ has a root in $I_{i}$.

Thus $\Delta S$ has at least $n$ roots; however, its degree is less than $n$, so $\Delta S$ should be identically zero. This shows that $S(x)$ is a constant, say $S(x) \equiv \beta$.
Step 3. Notice that the polynomials $P(x)-\beta / 2$ and $Q(x)-\beta / 2$ are also block-similar and distinct. So we may replace the initial polynomials by these ones, thus reaching $P(x)=-Q(x)$.

Then $R(x)=2 P(x)$, so $P(x)$ has exactly one root in each of the segments $I_{i}, i=1,2, \ldots, n$. On the other hand, $P(x)$ and $-P(x)$ should attain the same number of positive values on $Z_{i}$. Since $k$ is odd, this means that $Z_{i}$ contains exactly one root of $P(x)$; moreover, this root should be at the center of $Z_{i}$, because $P(x)$ has the same number of positive and negative values on $Z_{i}$.

Thus we have found all $n$ roots of $P(x)$, so

$$
P(x)=c \prod_{i=1}^{n}(x-i k+\ell) \quad \text { for some } c \in \mathbb{R} \backslash\{0\}
$$

where $\ell=(k-1) / 2$. It remains to notice that for every $t \in Z_{1} \backslash\{1\}$ we have

$$
|P(t)|=|c| \cdot|t-\ell-1| \cdot \prod_{i=2}^{n}|t-i k+\ell|<|c| \cdot \ell \cdot \prod_{i=2}^{n}|1-i k+\ell|=|P(1)|
$$

so $P(1) \neq-P(t)$ for all $t \in Z_{1}$. This shows that $P(x)$ is not block-similar to $-P(x)$. The final contradiction.

Comment 2. One may merge Steps 1 and 2 in the following manner. As above, we set $R(x)=$ $P(x)-Q(x)$ and $S(x)=P(x)+Q(x)$.

We aim to prove that the polynomial $S(x)=2 P(x)-R(x)=2 Q(x)+R(x)$ is constant. Since the degrees of $R(x)$ and $S(x)$ do not exceed $n$, it suffices to show that the total number of roots of $R(x)$ and $\Delta S(x)=S(x)-S(x-1)$ is at least $2 n$. For this purpose, we prove the following claim.
Claim. For every $i=1,2, \ldots, n$, either each of $R$ and $\Delta S$ has a root in $I_{i}$, or $R$ has at least two roots in $I_{i}$.
Proof. Fix any $i \in\{1,2, \ldots, n\}$. Let $r \in Z_{i}$ be a point such that $|R(r)|=\max _{x \in Z_{i}}|R(x)|$; we may assume that $R(r)>0$. Next, let $p^{-}, q^{+} \in I_{i}$ be some points such that $P\left(p^{-}\right)=\min _{x \in Z_{i}} P(x)$ and $Q\left(q^{+}\right)=\max _{x \in Z_{i}} Q(x)$. Notice that $P\left(p^{-}\right) \leqslant Q(r)<P(r)$ and $Q\left(q^{+}\right) \geqslant P(r)>Q(r)$, so $r$ is different from $p^{-}$and $q^{+}$.

Without loss of generality, we may assume that $p^{-}<r$. Then we have $R\left(p^{-}\right)=P\left(p^{-}\right)-Q\left(p^{-}\right) \leqslant$ $0<R(r)$, so $R(x)$ has a root in $\left[p^{-}, r\right)$. If $q^{+}>r$, then, similarly, $R\left(q^{+}\right) \leqslant 0<R(r)$, and $R(x)$ also has a root in $\left(r, q^{+}\right]$; so $R(x)$ has two roots in $I_{i}$, as required.

In the remaining case we have $q^{+}<r$; it suffices now to show that in this case $\Delta S$ has a root in $I_{i}$. Since $P\left(p^{-}\right) \leqslant Q(r)$ and $\left|R\left(p^{-}\right)\right| \leqslant R(r)$, we have $S\left(p^{-}\right)=2 P\left(p^{-}\right)-R\left(p^{-}\right) \leqslant 2 Q(r)+R(r)=S(r)$. Similarly, we get $S\left(q^{+}\right)=2 Q\left(q^{+}\right)+R\left(q^{+}\right) \geqslant 2 P(r)-R(r)=S(r)$. Therefore, the sequence of values of $S$ on $Z_{i}$ is neither strictly increasing nor strictly decreasing, which shows that $\Delta S$ has a root in $I_{i}$.

Comment 3. After finding the relation $P(x)-Q(x)=\alpha(T(x)-T(x-1))$ from Solution 1, one may also follow the approach presented in Solution 2. Knowledge of the difference of polynomials may simplify some steps; e.g., it is clear now that $P(x)-Q(x)$ has exactly one root in each of the segments $I_{i}$.

## Combinatorics

C1. In Lineland there are $n \geqslant 1$ towns, arranged along a road running from left to right. Each town has a left bulldozer (put to the left of the town and facing left) and a right bulldozer (put to the right of the town and facing right). The sizes of the $2 n$ bulldozers are distinct. Every time when a right and a left bulldozer confront each other, the larger bulldozer pushes the smaller one off the road. On the other hand, the bulldozers are quite unprotected at their rears; so, if a bulldozer reaches the rear-end of another one, the first one pushes the second one off the road, regardless of their sizes.

Let $A$ and $B$ be two towns, with $B$ being to the right of $A$. We say that town $A$ can sweep town $B$ away if the right bulldozer of $A$ can move over to $B$ pushing off all bulldozers it meets. Similarly, $B$ can sweep $A$ away if the left bulldozer of $B$ can move to $A$ pushing off all bulldozers of all towns on its way.

Prove that there is exactly one town which cannot be swept away by any other one.
(Estonia)
Solution 1. Let $T_{1}, T_{2}, \ldots, T_{n}$ be the towns enumerated from left to right. Observe first that, if town $T_{i}$ can sweep away town $T_{j}$, then $T_{i}$ also can sweep away every town located between $T_{i}$ and $T_{j}$.

We prove the problem statement by strong induction on $n$. The base case $n=1$ is trivial.
For the induction step, we first observe that the left bulldozer in $T_{1}$ and the right bulldozer in $T_{n}$ are completely useless, so we may forget them forever. Among the other $2 n-2$ bulldozers, we choose the largest one. Without loss of generality, it is the right bulldozer of some town $T_{k}$ with $k<n$.

Surely, with this large bulldozer $T_{k}$ can sweep away all the towns to the right of it. Moreover, none of these towns can sweep $T_{k}$ away; so they also cannot sweep away any town to the left of $T_{k}$. Thus, if we remove the towns $T_{k+1}, T_{k+2}, \ldots, T_{n}$, none of the remaining towns would change its status of being (un)sweepable away by the others.

Applying the induction hypothesis to the remaining towns, we find a unique town among $T_{1}, T_{2}, \ldots, T_{k}$ which cannot be swept away. By the above reasons, it is also the unique such town in the initial situation. Thus the induction step is established.

Solution 2. We start with the same enumeration and the same observation as in Solution 1. We also denote by $\ell_{i}$ and $r_{i}$ the sizes of the left and the right bulldozers belonging to $T_{i}$, respectively. One may easily see that no two towns $T_{i}$ and $T_{j}$ with $i<j$ can sweep each other away, for this would yield $r_{i}>\ell_{j}>r_{i}$.

Clearly, there is no town which can sweep $T_{n}$ away from the right. Then we may choose the leftmost town $T_{k}$ which cannot be swept away from the right. One can observe now that no town $T_{i}$ with $i>k$ may sweep away some town $T_{j}$ with $j<k$, for otherwise $T_{i}$ would be able to sweep $T_{k}$ away as well.

Now we prove two claims, showing together that $T_{k}$ is the unique town which cannot be swept away, and thus establishing the problem statement.
Claim 1. $T_{k}$ also cannot be swept away from the left.
Proof. Let $T_{m}$ be some town to the left of $T_{k}$. By the choice of $T_{k}$, town $T_{m}$ can be swept away from the right by some town $T_{p}$ with $p>m$. As we have already observed, $p$ cannot be greater than $k$. On the other hand, $T_{m}$ cannot sweep $T_{p}$ away, so a fortiori it cannot sweep $T_{k}$ away.

Claim 2. Any town $T_{m}$ with $m \neq k$ can be swept away by some other town.

Proof. If $m<k$, then $T_{m}$ can be swept away from the right due to the choice of $T_{k}$. In the remaining case we have $m>k$.

Let $T_{p}$ be a town among $T_{k}, T_{k+1}, \ldots, T_{m-1}$ having the largest right bulldozer. We claim that $T_{p}$ can sweep $T_{m}$ away. If this is not the case, then $r_{p}<\ell_{q}$ for some $q$ with $p<q \leqslant m$. But this means that $\ell_{q}$ is greater than all the numbers $r_{i}$ with $k \leqslant i \leqslant m-1$, so $T_{q}$ can sweep $T_{k}$ away. This contradicts the choice of $T_{k}$.

Comment 1. One may employ the same ideas within the inductive approach. Here we sketch such a solution.

Assume that the problem statement holds for the collection of towns $T_{1}, T_{2}, \ldots, T_{n-1}$, so that there is a unique town $T_{i}$ among them which cannot be swept away by any other of them. Thus we need to prove that in the full collection $T_{1}, T_{2}, \ldots, T_{n}$, exactly one of the towns $T_{i}$ and $T_{n}$ cannot be swept away.

If $T_{n}$ cannot sweep $T_{i}$ away, then it remains to prove that $T_{n}$ can be swept away by some other town. This can be established as in the second paragraph of the proof of Claim 2.

If $T_{n}$ can sweep $T_{i}$ away, then it remains to show that $T_{n}$ cannot be swept away by any other town. Since $T_{n}$ can sweep $T_{i}$ away, it also can sweep all the towns $T_{i}, T_{i+1}, \ldots, T_{n-1}$ away, so $T_{n}$ cannot be swept away by any of those. On the other hand, none of the remaining towns $T_{1}, T_{2}, \ldots, T_{i-1}$ can sweep $T_{i}$ away, so that they cannot sweep $T_{n}$ away as well.

Comment 2. Here we sketch yet another inductive approach. Assume that $n>1$. Firstly, we find a town which can be swept away by each of its neighbors (each town has two neighbors, except for the bordering ones each of which has one); we call such town a loser. Such a town exists, because there are $n-1$ pairs of neighboring towns, and in each of them there is only one which can sweep the other away; so there exists a town which is a winner in none of these pairs.

Notice that a loser can be swept away, but it cannot sweep any other town away (due to its neighbors' protection). Now we remove a loser, and suggest its left bulldozer to its right neighbor (if it exists), and its right bulldozer to a left one (if it exists). Surely, a town accepts a suggestion if a suggested bulldozer is larger than the town's one of the same orientation.

Notice that suggested bulldozers are useless in attack (by the definition of a loser), but may serve for defensive purposes. Moreover, each suggested bulldozer's protection works for the same pairs of remaining towns as before the removal.

By the induction hypothesis, the new configuration contains exactly one town which cannot be swept away. The arguments above show that the initial one also satisfies this property.

Solution 3. We separately prove that $(i)$ there exists a town which cannot be swept away, and that (ii) there is at most one such town. We also make use of the two observations from the previous solutions.
To prove $(i)$, assume contrariwise that every town can be swept away. Let $t_{1}$ be the leftmost town; next, for every $k=1,2, \ldots$ we inductively choose $t_{k+1}$ to be some town which can sweep $t_{k}$ away. Now we claim that for every $k=1,2, \ldots$, the town $t_{k+1}$ is to the right of $t_{k}$; this leads to the contradiction, since the number of towns is finite.

Induction on $k$. The base case $k=1$ is clear due to the choice of $t_{1}$. Assume now that for all $j$ with $1 \leqslant j<k$, the town $t_{j+1}$ is to the right of $t_{j}$. Suppose that $t_{k+1}$ is situated to the left of $t_{k}$; then it lies between $t_{j}$ and $t_{j+1}$ (possibly coinciding with $t_{j}$ ) for some $j<k$. Therefore, $t_{k+1}$ can be swept away by $t_{j+1}$, which shows that it cannot sweep $t_{j+1}$ away - so $t_{k+1}$ also cannot sweep $t_{k}$ away. This contradiction proves the induction step.

To prove (ii), we also argue indirectly and choose two towns $A$ and $B$ neither of which can be swept away, with $A$ being to the left of $B$. Consider the largest bulldozer $b$ between them (taking into consideration the right bulldozer of $A$ and the left bulldozer of $B$ ). Without loss of generality, $b$ is a left bulldozer; then it is situated in some town to the right of $A$, and this town may sweep $A$ away since nothing prevents it from doing that. A contradiction.

Comment 3. The Problem Selection Committee decided to reformulate this problem. The original formulation was as follows.

Let $n$ be a positive integer. There are $n$ cards in a deck, enumerated from bottom to top with numbers $1,2, \ldots, n$. For each $i=1,2, \ldots, n$, an even number $a_{i}$ is printed on the lower side and an odd number $b_{i}$ is printed on the upper side of the $i^{\text {th }}$ card. We say that the $i^{\text {th }}$ card opens the $j^{\text {th }}$ card, if $i<j$ and $b_{i}<a_{k}$ for every $k=i+1, i+2, \ldots, j$. Similarly, we say that the $i^{\text {th }}$ card closes the $j^{\text {th }}$ card, if $i>j$ and $a_{i}<b_{k}$ for every $k=i-1, i-2, \ldots, j$. Prove that the deck contains exactly one card which is neither opened nor closed by any other card.

C2. Let $\mathcal{V}$ be a finite set of points in the plane. We say that $\mathcal{V}$ is balanced if for any two distinct points $A, B \in \mathcal{V}$, there exists a point $C \in \mathcal{V}$ such that $A C=B C$. We say that $\mathcal{V}$ is center-free if for any distinct points $A, B, C \in \mathcal{V}$, there does not exist a point $P \in \mathcal{V}$ such that $P A=P B=P C$.
(a) Show that for all $n \geqslant 3$, there exists a balanced set consisting of $n$ points.
(b) For which $n \geqslant 3$ does there exist a balanced, center-free set consisting of $n$ points?
(Netherlands)
Answer for part (b). All odd integers $n \geqslant 3$.

## Solution.

Part ( $\boldsymbol{a}$ ). Assume that $n$ is odd. Consider a regular $n$-gon. Label the vertices of the $n$-gon as $A_{1}, A_{2}, \ldots, A_{n}$ in counter-clockwise order, and set $\mathcal{V}=\left\{A_{1}, \ldots, A_{n}\right\}$. We check that $\mathcal{V}$ is balanced. For any two distinct vertices $A_{i}$ and $A_{j}$, let $k \in\{1,2, \ldots, n\}$ be the solution of $2 k \equiv i+j(\bmod n)$. Then, since $k-i \equiv j-k(\bmod n)$, we have $A_{i} A_{k}=A_{j} A_{k}$, as required.

Now assume that $n$ is even. Consider a regular ( $3 n-6$ )-gon, and let $O$ be its circumcenter. Again, label its vertices as $A_{1}, \ldots, A_{3 n-6}$ in counter-clockwise order, and choose $\mathcal{V}=$ $\left\{O, A_{1}, A_{2}, \ldots, A_{n-1}\right\}$. We check that $\mathcal{V}$ is balanced. For any two distinct vertices $A_{i}$ and $A_{j}$, we always have $O A_{i}=O A_{j}$. We now consider the vertices $O$ and $A_{i}$. First note that the triangle $O A_{i} A_{n / 2-1+i}$ is equilateral for all $i \leqslant \frac{n}{2}$. Hence, if $i \leqslant \frac{n}{2}$, then we have $O A_{n / 2-1+i}=A_{i} A_{n / 2-1+i}$; otherwise, if $i>\frac{n}{2}$, then we have $O A_{i-n / 2+1}=A_{i} A_{i-n / 2+1}$. This completes the proof.

An example of such a construction when $n=10$ is shown in Figure 1.


Figure 1


Figure 2

Comment (a). There are many ways to construct an example by placing equilateral triangles in a circle. Here we present one general method.

Let $O$ be the center of a circle and let $A_{1}, B_{1}, \ldots, A_{k}, B_{k}$ be distinct points on the circle such that the triangle $O A_{i} B_{i}$ is equilateral for each $i$. Then $\mathcal{V}=\left\{O, A_{1}, B_{1}, \ldots, A_{k}, B_{k}\right\}$ is balanced. To construct a set of even cardinality, put extra points $C, D, E$ on the circle such that triangles $O C D$ and $O D E$ are equilateral (see Figure 2). Then $\mathcal{V}=\left\{O, A_{1}, B_{1}, \ldots, A_{k}, B_{k}, C, D, E\right\}$ is balanced.

Part (b). We now show that there exists a balanced, center-free set containing $n$ points for all odd $n \geqslant 3$, and that one does not exist for any even $n \geqslant 3$.

If $n$ is odd, then let $\mathcal{V}$ be the set of vertices of a regular $n$-gon. We have shown in part ( $a$ ) that $\mathcal{V}$ is balanced. We claim that $\mathcal{V}$ is also center-free. Indeed, if $P$ is a point such that
$P A=P B=P C$ for some three distinct vertices $A, B$ and $C$, then $P$ is the circumcenter of the $n$-gon, which is not contained in $\mathcal{V}$.

Now suppose that $\mathcal{V}$ is a balanced, center-free set of even cardinality $n$. We will derive a contradiction. For a pair of distinct points $A, B \in \mathcal{V}$, we say that a point $C \in \mathcal{V}$ is associated with the pair $\{A, B\}$ if $A C=B C$. Since there are $\frac{n(n-1)}{2}$ pairs of points, there exists a point $P \in \mathcal{V}$ which is associated with at least $\left\lceil\frac{n(n-1)}{2} / n\right\rceil=\frac{n}{2}$ pairs. Note that none of these $\frac{n}{2}$ pairs can contain $P$, so that the union of these $\frac{n}{2}$ pairs consists of at most $n-1$ points. Hence there exist two such pairs that share a point. Let these two pairs be $\{A, B\}$ and $\{A, C\}$. Then $P A=P B=P C$, which is a contradiction.

Comment (b). We can rephrase the argument in graph theoretic terms as follows. Let $\mathcal{V}$ be a balanced, center-free set consisting of $n$ points. For any pair of distinct vertices $A, B \in \mathcal{V}$ and for any $C \in \mathcal{V}$ such that $A C=B C$, draw directed edges $A \rightarrow C$ and $B \rightarrow C$. Then all pairs of vertices generate altogether at least $n(n-1)$ directed edges; since the set is center-free, these edges are distinct. So we must obtain a graph in which any two vertices are connected in both directions. Now, each vertex has exactly $n-1$ incoming edges, which means that $n-1$ is even. Hence $n$ is odd.

C3. For a finite set $A$ of positive integers, we call a partition of $A$ into two disjoint nonempty subsets $A_{1}$ and $A_{2}$ good if the least common multiple of the elements in $A_{1}$ is equal to the greatest common divisor of the elements in $A_{2}$. Determine the minimum value of $n$ such that there exists a set of $n$ positive integers with exactly 2015 good partitions.
(Ukraine)
Answer. 3024.
Solution. Let $A=\left\{a_{1}, a_{2}, \ldots, a_{n}\right\}$, where $a_{1}<a_{2}<\cdots<a_{n}$. For a finite nonempty set $B$ of positive integers, denote by $\operatorname{lcm} B$ and $\operatorname{gcd} B$ the least common multiple and the greatest common divisor of the elements in $B$, respectively.

Consider any good partition $\left(A_{1}, A_{2}\right)$ of $A$. By definition, $\operatorname{lcm} A_{1}=d=\operatorname{gcd} A_{2}$ for some positive integer $d$. For any $a_{i} \in A_{1}$ and $a_{j} \in A_{2}$, we have $a_{i} \leqslant d \leqslant a_{j}$. Therefore, we have $A_{1}=\left\{a_{1}, a_{2}, \ldots, a_{k}\right\}$ and $A_{2}=\left\{a_{k+1}, a_{k+2}, \ldots, a_{n}\right\}$ for some $k$ with $1 \leqslant k<n$. Hence, each good partition is determined by an element $a_{k}$, where $1 \leqslant k<n$. We call such $a_{k}$ partitioning.

It is convenient now to define $\ell_{k}=\operatorname{lcm}\left(a_{1}, a_{2}, \ldots, a_{k}\right)$ and $g_{k}=\operatorname{gcd}\left(a_{k+1}, a_{k+2}, \ldots, a_{n}\right)$ for $1 \leqslant k \leqslant n-1$. So $a_{k}$ is partitioning exactly when $\ell_{k}=g_{k}$.

We proceed by proving some properties of partitioning elements, using the following claim. Claim. If $a_{k-1}$ and $a_{k}$ are partitioning where $2 \leqslant k \leqslant n-1$, then $g_{k-1}=g_{k}=a_{k}$.
Proof. Assume that $a_{k-1}$ and $a_{k}$ are partitioning. Since $\ell_{k-1}=g_{k-1}$, we have $\ell_{k-1} \mid a_{k}$. Therefore, $g_{k}=\ell_{k}=\operatorname{lcm}\left(\ell_{k-1}, a_{k}\right)=a_{k}$, and $g_{k-1}=\operatorname{gcd}\left(a_{k}, g_{k}\right)=a_{k}$, as desired.

Property 1. For every $k=2,3, \ldots, n-2$, at least one of $a_{k-1}, a_{k}$, and $a_{k+1}$ is not partitioning. Proof. Suppose, to the contrary, that all three numbers $a_{k-1}, a_{k}$, and $a_{k+1}$ are partitioning. The claim yields that $a_{k+1}=g_{k}=a_{k}$, a contradiction.
Property 2. The elements $a_{1}$ and $a_{2}$ cannot be simultaneously partitioning. Also, $a_{n-2}$ and $\overline{a_{n-1} \text { cannot be simultaneously partitioning }}$
Proof. Assume that $a_{1}$ and $a_{2}$ are partitioning. By the claim, it follows that $a_{2}=g_{1}=\ell_{1}=$ $\operatorname{lcm}\left(a_{1}\right)=a_{1}$, a contradiction.

Similarly, assume that $a_{n-2}$ and $a_{n-1}$ are partitioning. The claim yields that $a_{n-1}=g_{n-1}=$ $\operatorname{gcd}\left(a_{n}\right)=a_{n}$, a contradiction.

Now let $A$ be an $n$-element set with exactly 2015 good partitions. Clearly, we have $n \geqslant 5$. Using Property 2, we find that there is at most one partitioning element in each of $\left\{a_{1}, a_{2}\right\}$ and $\left\{a_{n-2}, a_{n-1}\right\}$. By Property 1 , there are at least $\left\lfloor\frac{n-5}{3}\right\rfloor$ non-partitioning elements in $\left\{a_{3}, a_{4}, \ldots, a_{n-3}\right\}$. Therefore, there are at most $(n-1)-2-\left\lfloor\frac{n-5}{3}\right\rfloor=\left\lceil\frac{2(n-2)}{3}\right\rceil$ partitioning elements in $A$. Thus, $\left\lceil\frac{2(n-2)}{3}\right\rceil \geqslant 2015$, which implies that $n \geqslant 3024$.

Finally, we show that there exists a set of 3024 positive integers with exactly 2015 partitioning elements. Indeed, in the set $A=\left\{2 \cdot 6^{i}, 3 \cdot 6^{i}, 6^{i+1} \mid 0 \leqslant i \leqslant 1007\right\}$, each element of the form $3 \cdot 6^{i}$ or $6^{i}$, except $6^{1008}$, is partitioning.

Therefore, the minimum possible value of $n$ is 3024 .
Comment. Here we will work out the general case when 2015 is replaced by an arbitrary positive integer $m$. Note that the bound $\left\lceil\frac{2(n-2)}{3}\right\rceil \geqslant m$ obtained in the solution is, in fact, true for any positive integers $m$ and $n$. Using this bound, one can find that $n \geqslant\left\lceil\frac{3 m}{2}\right\rceil+1$.

To show that the bound is sharp, one constructs a set of $\left\lceil\frac{3 m}{2}\right\rceil+1$ elements with exactly $m$ good partitions. Indeed, the minimum is attained on the set $\left\{6^{i}, 2 \cdot 6^{i}, 3 \cdot 6^{i} \mid 0 \leqslant i \leqslant t-1\right\} \cup\left\{6^{t}\right\}$ for every even $m=2 t$, and $\left\{2 \cdot 6^{i}, 3 \cdot 6^{i}, 6^{i+1} \mid 0 \leqslant i \leqslant t-1\right\}$ for every odd $m=2 t-1$.
$\mathbf{C 4}$. Let $n$ be a positive integer. Two players $A$ and $B$ play a game in which they take turns choosing positive integers $k \leqslant n$. The rules of the game are:
(i) A player cannot choose a number that has been chosen by either player on any previous turn.
(ii) A player cannot choose a number consecutive to any of those the player has already chosen on any previous turn.
(iii) The game is a draw if all numbers have been chosen; otherwise the player who cannot choose a number anymore loses the game.

The player $A$ takes the first turn. Determine the outcome of the game, assuming that both players use optimal strategies.

(Finland)

Answer. The game ends in a draw when $n=1,2,4,6$; otherwise $B$ wins.
Solution. For brevity, we denote by $[n]$ the set $\{1,2, \ldots, n\}$.
Firstly, we show that $B$ wins whenever $n \neq 1,2,4,6$. For this purpose, we provide a strategy which guarantees that $B$ can always make a move after $A$ 's move, and also guarantees that the game does not end in a draw.

We begin with an important observation.
Lemma. Suppose that $B$ 's first pick is $n$ and that $A$ has made the $k^{\text {th }}$ move where $k \geqslant 2$. Then $B$ can also make the $k^{\text {th }}$ move.
Proof. Let $\mathcal{S}$ be the set of the first $k$ numbers chosen by $A$. Since $\mathcal{S}$ does not contain consecutive integers, we see that the set $[n] \backslash \mathcal{S}$ consists of $k$ "contiguous components" if $1 \in \mathcal{S}$, and $k+1$ components otherwise. Since $B$ has chosen only $k-1$ numbers, there is at least one component of $[n] \backslash \mathcal{S}$ consisting of numbers not yet picked by $B$. Hence, $B$ can choose a number from this component.

We will now describe a winning strategy for $B$, when $n \neq 1,2,4,6$. By symmetry, we may assume that $A$ 's first choice is a number not exceeding $\frac{n+1}{2}$. So $B$ can pick the number $n$ in $B$ 's first turn. We now consider two cases.

Case 1. $n$ is odd and $n \geqslant 3$. The only way the game ends in a draw is that $A$ eventually picks all the odd numbers from the set $[n]$. However, $B$ has already chosen $n$, so this cannot happen. Thus $B$ can continue to apply the lemma until $A$ cannot make a move.

Case 2. $n$ is even and $n \geqslant 8$. Since $B$ has picked $n$, the game is a draw only if $A$ can eventually choose all the odd numbers from the set $[n-1]$. So $B$ picks a number from the set $\{1,3,5, \ldots, n-3\}$ not already chosen by $A$, on $B$ 's second move. This is possible since the set consists of $\frac{n-2}{2} \geqslant 3$ numbers and $A$ has chosen only 2 numbers. Hereafter $B$ can apply the lemma until $A$ cannot make a move.

Hence, in both cases $A$ loses.
We are left with the cases $n=1,2,4,6$. The game is trivially a draw when $n=1,2$. When $n=4, A$ has to first pick 1 to avoid losing. Similarly, $B$ has to choose 4 as well. It then follows that the game ends in a draw.

When $n=6, B$ gets at least a draw by the lemma or by using a mirror strategy. On the other hand, $A$ may also get at least a draw in the following way. In the first turn, $A$ chooses 1 . After $B$ 's response by a number $b, A$ finds a neighbor $c$ of $b$ which differs from 1 and 2 , and reserves $c$ for $A$ 's third move. Now, clearly $A$ can make the second move by choosing a number different from $1,2, c-1, c, c+1$. Therefore $A$ will not lose.

Comment 1. We present some explicit winning strategies for $B$.
We start with the case $n$ is odd and $n \geqslant 3 . B$ starts by picking $n$ in the first turn. On the $k^{\text {th }}$ move for $k \geqslant 2, B$ chooses the number exactly 1 less than $A^{\prime}$ 's $k^{\text {th }}$ pick. The only special case is when $A$ 's $k^{\text {th }}$ choice is 1 . In this situation, $A$ 's first pick was a number $a>1$ and $B$ can respond by choosing $a-1$ on the $k^{\text {th }}$ move instead.

We now give an alternative winning strategy in the case $n$ is even and $n \geqslant 8$. We first present a winning strategy for the case when $A$ 's first pick is 1 . We consider two cases depending on $A$ 's second move.

Case 1. A's second pick is 3 . Then $B$ chooses $n-3$ on the second move. On the $k^{\text {th }}$ move, $B$ chooses the number exactly 1 less than $A$ 's $k^{\text {th }}$ pick except that $B$ chooses 2 if $A$ 's $k^{\text {th }}$ pick is $n-2$ or $n-1$.

Case 2. A's second pick is $a>3$. Then $B$ chooses $a-2$ on the second move. Afterwards on the $k^{\text {th }}$ move, $B$ picks the number exactly 1 less than $A^{\prime}$ 's $k^{\text {th }}$ pick.

One may easily see that this strategy guarantees $B$ 's victory, when $A$ 's first pick is 1 .
The following claim shows how to extend the strategy to the general case.
Claim. Assume that $B$ has an explicit strategy leading to a victory after $A$ picks 1 on the first move. Then $B$ also has an explicit strategy leading to a victory after any first moves of $A$.
Proof. Let $S$ be an optimal strategy of $B$ after $A$ picks 1 on the first move. Assume that $A$ picks some number $a>1$ on this move; we show how $B$ can make use of $S$ in order to win in this case.

In parallel to the real play, $B$ starts an imaginary play. The positions in these plays differ by flipping the segment $[1, a]$; so, if a player chooses some number $x$ in the real play, then the same player chooses a number $x$ or $a+1-x$ in the imaginary play, depending on whether $x>a$ or $x \leqslant a$. Thus $A$ 's first pick in the imaginary play is 1 .

Clearly, a number is chosen in the real play exactly if the corresponding number is chosen in the imaginary one. Next, if an unchosen number is neighboring to one chosen by $A$ in the imaginary play, then the corresponding number also has this property in the real play, so $A$ also cannot choose it. One can easily see that a similar statement with real and imaginary plays interchanged holds for $B$ instead of $A$.

Thus, when $A$ makes some move in the real play, $B$ may imagine the corresponding legal move in the imaginary one. Then $B$ chooses the response according to $S$ in the imaginary game and makes the corresponding legal move in the real one. Acting so, $B$ wins the imaginary game, thus $B$ will also win the real one.

Hence, $B$ has a winning strategy for all even $n$ greater or equal to 8 .
Notice that the claim can also be used to simplify the argument when $n$ is odd.
Comment 2. One may also employ symmetry when $n$ is odd. In particular, $B$ could use a mirror strategy. However, additional ideas are required to modify the strategy after $A$ picks $\frac{n+1}{2}$.

C5. Consider an infinite sequence $a_{1}, a_{2}, \ldots$ of positive integers with $a_{i} \leqslant 2015$ for all $i \geqslant 1$. Suppose that for any two distinct indices $i$ and $j$ we have $i+a_{i} \neq j+a_{j}$.

Prove that there exist two positive integers $b$ and $N$ such that

$$
\left|\sum_{i=m+1}^{n}\left(a_{i}-b\right)\right| \leqslant 1007^{2}
$$

whenever $n>m \geqslant N$.
(Australia)
Solution 1. We visualize the set of positive integers as a sequence of points. For each $n$ we draw an arrow emerging from $n$ that points to $n+a_{n}$; so the length of this arrow is $a_{n}$. Due to the condition that $m+a_{m} \neq n+a_{n}$ for $m \neq n$, each positive integer receives at most one arrow. There are some positive integers, such as 1 , that receive no arrows; these will be referred to as starting points in the sequel. When one starts at any of the starting points and keeps following the arrows, one is led to an infinite path, called its ray, that visits a strictly increasing sequence of positive integers. Since the length of any arrow is at most 2015, such a ray, say with starting point $s$, meets every interval of the form [ $n, n+2014]$ with $n \geqslant s$ at least once.

Suppose for the sake of contradiction that there would be at least 2016 starting points. Then we could take an integer $n$ that is larger than the first 2016 starting points. But now the interval $[n, n+2014]$ must be met by at least 2016 rays in distinct points, which is absurd. We have thereby shown that the number $b$ of starting points satisfies $1 \leqslant b \leqslant 2015$. Let $N$ denote any integer that is larger than all starting points. We contend that $b$ and $N$ are as required.

To see this, let any two integers $m$ and $n$ with $n>m \geqslant N$ be given. The sum $\sum_{i=m+1}^{n} a_{i}$ gives the total length of the arrows emerging from $m+1, \ldots, n$. Taken together, these arrows form $b$ subpaths of our rays, some of which may be empty. Now on each ray we look at the first number that is larger than $m$; let $x_{1}, \ldots, x_{b}$ denote these numbers, and let $y_{1}, \ldots, y_{b}$ enumerate in corresponding order the numbers defined similarly with respect to $n$. Then the list of differences $y_{1}-x_{1}, \ldots, y_{b}-x_{b}$ consists of the lengths of these paths and possibly some zeros corresponding to empty paths. Consequently, we obtain

$$
\sum_{i=m+1}^{n} a_{i}=\sum_{j=1}^{b}\left(y_{j}-x_{j}\right)
$$

whence

$$
\sum_{i=m+1}^{n}\left(a_{i}-b\right)=\sum_{j=1}^{b}\left(y_{j}-n\right)-\sum_{j=1}^{b}\left(x_{j}-m\right) .
$$

Now each of the $b$ rays meets the interval $[m+1, m+2015]$ at some point and thus $x_{1}-$ $m, \ldots, x_{b}-m$ are $b$ distinct members of the set $\{1,2, \ldots, 2015\}$. Moreover, since $m+1$ is not a starting point, it must belong to some ray; so 1 has to appear among these numbers, wherefore

$$
1+\sum_{j=1}^{b-1}(j+1) \leqslant \sum_{j=1}^{b}\left(x_{j}-m\right) \leqslant 1+\sum_{j=1}^{b-1}(2016-b+j) .
$$

The same argument applied to $n$ and $y_{1}, \ldots, y_{b}$ yields

$$
1+\sum_{j=1}^{b-1}(j+1) \leqslant \sum_{j=1}^{b}\left(y_{j}-n\right) \leqslant 1+\sum_{j=1}^{b-1}(2016-b+j) .
$$

So altogether we get

$$
\begin{aligned}
\left|\sum_{i=m+1}^{n}\left(a_{i}-b\right)\right| & \leqslant \sum_{j=1}^{b-1}((2016-b+j)-(j+1))=(b-1)(2015-b) \\
& \leqslant\left(\frac{(b-1)+(2015-b)}{2}\right)^{2}=1007^{2}
\end{aligned}
$$

as desired.
Solution 2. Set $s_{n}=n+a_{n}$ for all positive integers $n$. By our assumptions, we have

$$
n+1 \leqslant s_{n} \leqslant n+2015
$$

for all $n \in \mathbb{Z}_{>0}$. The members of the sequence $s_{1}, s_{2}, \ldots$ are distinct. We shall investigate the set

$$
M=\mathbb{Z}_{>0} \backslash\left\{s_{1}, s_{2}, \ldots\right\}
$$

Claim. At most 2015 numbers belong to $M$.
Proof. Otherwise let $m_{1}<m_{2}<\cdots<m_{2016}$ be any 2016 distinct elements from $M$. For $n=m_{2016}$ we have

$$
\left\{s_{1}, \ldots, s_{n}\right\} \cup\left\{m_{1}, \ldots, m_{2016}\right\} \subseteq\{1,2, \ldots, n+2015\},
$$

where on the left-hand side we have a disjoint union containing altogether $n+2016$ elements. But the set on the right-hand side has only $n+2015$ elements. This contradiction proves our claim.

Now we work towards proving that the positive integers $b=|M|$ and $N=\max (M)$ are as required. Recall that we have just shown $b \leqslant 2015$.

Let us consider any integer $r \geqslant N$. As in the proof of the above claim, we see that

$$
\begin{equation*}
B_{r}=M \cup\left\{s_{1}, \ldots, s_{r}\right\} \tag{1}
\end{equation*}
$$

is a subset of $[1, r+2015] \cap \mathbb{Z}$ with precisely $b+r$ elements. Due to the definitions of $M$ and $N$, we also know $[1, r+1] \cap \mathbb{Z} \subseteq B_{r}$. It follows that there is a set $C_{r} \subseteq\{1,2, \ldots, 2014\}$ with $\left|C_{r}\right|=b-1$ and

$$
\begin{equation*}
B_{r}=([1, r+1] \cap \mathbb{Z}) \cup\left\{r+1+x \mid x \in C_{r}\right\} \tag{2}
\end{equation*}
$$

For any finite set of integers $J$ we denote the sum of its elements by $\sum J$. Now the equations (1) and (2) give rise to two ways of computing $\sum B_{r}$ and the comparison of both methods leads to

$$
\sum M+\sum_{i=1}^{r} s_{i}=\sum_{i=1}^{r} i+b(r+1)+\sum C_{r}
$$

or in other words to

$$
\begin{equation*}
\sum M+\sum_{i=1}^{r}\left(a_{i}-b\right)=b+\sum C_{r} . \tag{3}
\end{equation*}
$$

After this preparation, we consider any two integers $m$ and $n$ with $n>m \geqslant N$. Plugging $r=n$ and $r=m$ into (3) and subtracting the estimates that result, we deduce

$$
\sum_{i=m+1}^{n}\left(a_{i}-b\right)=\sum C_{n}-\sum C_{m}
$$

Since $C_{n}$ and $C_{m}$ are subsets of $\{1,2, \ldots, 2014\}$ with $\left|C_{n}\right|=\left|C_{m}\right|=b-1$, it is clear that the absolute value of the right-hand side of the above inequality attains its largest possible value if either $C_{m}=\{1,2, \ldots, b-1\}$ and $C_{n}=\{2016-b, \ldots, 2014\}$, or the other way around. In these two cases we have

$$
\left|\sum C_{n}-\sum C_{m}\right|=(b-1)(2015-b)
$$

so in the general case we find

$$
\left|\sum_{i=m+1}^{n}\left(a_{i}-b\right)\right| \leqslant(b-1)(2015-b) \leqslant\left(\frac{(b-1)+(2015-b)}{2}\right)^{2}=1007^{2}
$$

as desired.

Comment. The sets $C_{n}$ may be visualized by means of the following process: Start with an empty blackboard. For $n \geqslant 1$, the following happens during the $n^{\text {th }}$ step. The number $a_{n}$ gets written on the blackboard, then all numbers currently on the blackboard are decreased by 1 , and finally all zeros that have arisen get swept away.

It is not hard to see that the numbers present on the blackboard after $n$ steps are distinct and form the set $C_{n}$. Moreover, it is possible to complete a solution based on this idea.

C6. Let $S$ be a nonempty set of positive integers. We say that a positive integer $n$ is clean if it has a unique representation as a sum of an odd number of distinct elements from $S$. Prove that there exist infinitely many positive integers that are not clean.

Solution 1. Define an odd (respectively, even) representation of $n$ to be a representation of $n$ as a sum of an odd (respectively, even) number of distinct elements of $S$. Let $\mathbb{Z}_{>0}$ denote the set of all positive integers.

Suppose, to the contrary, that there exist only finitely many positive integers that are not clean. Therefore, there exists a positive integer $N$ such that every integer $n>N$ has exactly one odd representation.

Clearly, $S$ is infinite. We now claim the following properties of odd and even representations.
Property 1. Any positive integer $n$ has at most one odd and at most one even representation.
Proof. We first show that every integer $n$ has at most one even representation. Since $S$ is infinite, there exists $x \in S$ such that $x>\max \{n, N\}$. Then, the number $n+x$ must be clean, and $x$ does not appear in any even representation of $n$. If $n$ has more than one even representation, then we obtain two distinct odd representations of $n+x$ by adding $x$ to the even representations of $n$, which is impossible. Therefore, $n$ can have at most one even representation.

Similarly, there exist two distinct elements $y, z \in S$ such that $y, z>\max \{n, N\}$. If $n$ has more than one odd representation, then we obtain two distinct odd representations of $n+y+z$ by adding $y$ and $z$ to the odd representations of $n$. This is again a contradiction.

Property 2. Fix $s \in S$. Suppose that a number $n>N$ has no even representation. Then $n+2 a s$ has an even representation containing $s$ for all integers $a \geqslant 1$.
Proof. It is sufficient to prove the following statement: If $n$ has no even representation without $s$, then $n+2 s$ has an even representation containing $s$ (and hence no even representation without $s$ by Property 1).

Notice that the odd representation of $n+s$ does not contain $s$; otherwise, we have an even representation of $n$ without $s$. Then, adding $s$ to this odd representation of $n+s$, we get that $n+2 s$ has an even representation containing $s$, as desired.

Property 3. Every sufficiently large integer has an even representation.
Proof. Fix any $s \in S$, and let $r$ be an arbitrary element in $\{1,2, \ldots, 2 s\}$. Then, Property 2 implies that the set $Z_{r}=\{r+2 a s: a \geqslant 0\}$ contains at most one number exceeding $N$ with no even representation. Therefore, $Z_{r}$ contains finitely many positive integers with no even representation, and so does $\mathbb{Z}_{>0}=\bigcup_{r=1}^{2 s} Z_{r}$.

In view of Properties 1 and 3, we may assume that $N$ is chosen such that every $n>N$ has exactly one odd and exactly one even representation. In particular, each element $s>N$ of $S$ has an even representation.

Property 4. For any $s, t \in S$ with $N<s<t$, the even representation of $t$ contains $s$.
Proof. Suppose the contrary. Then, $s+t$ has at least two odd representations: one obtained by adding $s$ to the even representation of $t$ and one obtained by adding $t$ to the even representation of $s$. Since the latter does not contain $s$, these two odd representations of $s+t$ are distinct, a contradiction.

Let $s_{1}<s_{2}<\cdots$ be all the elements of $S$, and set $\sigma_{n}=\sum_{i=1}^{n} s_{i}$ for each nonnegative integer $n$. Fix an integer $k$ such that $s_{k}>N$. Then, Property 4 implies that for every $i>k$ the even representation of $s_{i}$ contains all the numbers $s_{k}, s_{k+1}, \ldots, s_{i-1}$. Therefore,

$$
\begin{equation*}
s_{i}=s_{k}+s_{k+1}+\cdots+s_{i-1}+R_{i}=\sigma_{i-1}-\sigma_{k-1}+R_{i} \tag{1}
\end{equation*}
$$

where $R_{i}$ is a sum of some of $s_{1}, \ldots, s_{k-1}$. In particular, $0 \leqslant R_{i} \leqslant s_{1}+\cdots+s_{k-1}=\sigma_{k-1}$.

Let $j_{0}$ be an integer satisfying $j_{0}>k$ and $\sigma_{j_{0}}>2 \sigma_{k-1}$. Then (1) shows that, for every $j>j_{0}$,

$$
\begin{equation*}
s_{j+1} \geqslant \sigma_{j}-\sigma_{k-1}>\sigma_{j} / 2 . \tag{2}
\end{equation*}
$$

Next, let $p>j_{0}$ be an index such that $R_{p}=\min _{i>j_{0}} R_{i}$. Then,

$$
s_{p+1}=s_{k}+s_{k+1}+\cdots+s_{p}+R_{p+1}=\left(s_{p}-R_{p}\right)+s_{p}+R_{p+1} \geqslant 2 s_{p}
$$

Therefore, there is no element of $S$ larger than $s_{p}$ but smaller than $2 s_{p}$. It follows that the even representation $\tau$ of $2 s_{p}$ does not contain any element larger than $s_{p}$. On the other hand, inequality (2) yields $2 s_{p}>s_{1}+\cdots+s_{p-1}$, so $\tau$ must contain a term larger than $s_{p-1}$. Thus, it must contain $s_{p}$. After removing $s_{p}$ from $\tau$, we have that $s_{p}$ has an odd representation not containing $s_{p}$, which contradicts Property 1 since $s_{p}$ itself also forms an odd representation of $s_{p}$.

Solution 2. We will also use Property 1 from Solution 1.
We first define some terminology and notations used in this solution. Let $\mathbb{Z}_{\geqslant 0}$ denote the set of all nonnegative integers. All sums mentioned are regarded as sums of distinct elements of $S$. Moreover, a sum is called even or odd depending on the parity of the number of terms in it. All closed or open intervals refer to sets of all integers inside them, e.g., $[a, b]=\{x \in \mathbb{Z}: a \leqslant x \leqslant b\}$.

Again, let $s_{1}<s_{2}<\cdots$ be all elements of $S$, and denote $\sigma_{n}=\sum_{i=1}^{n} s_{i}$ for each positive integer $n$. Let $O_{n}$ (respectively, $E_{n}$ ) be the set of numbers representable as an odd (respectively, even) sum of elements of $\left\{s_{1}, \ldots, s_{n}\right\}$. Set $E=\bigcup_{n=1}^{\infty} E_{n}$ and $O=\bigcup_{n=1}^{\infty} O_{n}$. We assume that $0 \in E_{n}$ since 0 is representable as a sum of 0 terms.

We now proceed to our proof. Assume, to the contrary, that there exist only finitely many positive integers that are not clean and denote the number of non-clean positive integers by $m-1$. Clearly, $S$ is infinite. By Property 1 from Solution 1, every positive integer $n$ has at most one odd and at most one even representation.
Step 1. We estimate $s_{n+1}$ and $\sigma_{n+1}$.
Upper bounds: Property 1 yields $\left|O_{n}\right|=\left|E_{n}\right|=2^{n-1}$, so $\left|\left[1,2^{n-1}+m\right] \backslash O_{n}\right| \geqslant m$. Hence, there exists a clean integer $x_{n} \in\left[1,2^{n-1}+m\right] \backslash O_{n}$. The definition of $O_{n}$ then yields that the odd representation of $x_{n}$ contains a term larger than $s_{n}$. Therefore, $s_{n+1} \leqslant x_{n} \leqslant 2^{n-1}+m$ for every positive integer $n$. Moreover, since $s_{1}$ is the smallest clean number, we get $\sigma_{1}=s_{1} \leqslant m$. Then,

$$
\sigma_{n+1}=\sum_{i=2}^{n+1} s_{i}+s_{1} \leqslant \sum_{i=2}^{n+1}\left(2^{i-2}+m\right)+m=2^{n}-1+(n+1) m
$$

for every positive integer $n$. Notice that this estimate also holds for $n=0$.
Lower bounds: Since $O_{n+1} \subseteq\left[1, \sigma_{n+1}\right]$, we have $\sigma_{n+1} \geqslant\left|O_{n+1}\right|=2^{n}$ for all positive integers $n$. Then,

$$
s_{n+1}=\sigma_{n+1}-\sigma_{n} \geqslant 2^{n}-\left(2^{n-1}-1+n m\right)=2^{n-1}+1-n m
$$

for every positive integer $n$.
Combining the above inequalities, we have

$$
\begin{equation*}
2^{n-1}+1-n m \leqslant s_{n+1} \leqslant 2^{n-1}+m \quad \text { and } \quad 2^{n} \leqslant \sigma_{n+1} \leqslant 2^{n}-1+(n+1) m \tag{3}
\end{equation*}
$$

for every positive integer $n$.
Step 2. We prove Property 3 from Solution 1.
For every integer $x$ and set of integers $Y$, define $x \pm Y=\{x \pm y: y \in Y\}$.
In view of Property 1, we get

$$
E_{n+1}=E_{n} \sqcup\left(s_{n+1}+O_{n}\right) \quad \text { and } \quad O_{n+1}=O_{n} \sqcup\left(s_{n+1}+E_{n}\right),
$$

where $\sqcup$ denotes the disjoint union operator. Notice also that $s_{n+2} \geqslant 2^{n}+1-(n+1) m>$ $2^{n-1}-1+n m \geqslant \sigma_{n}$ for every sufficiently large $n$. We now claim the following.

Claim 1. $\left(\sigma_{n}-s_{n+1}, s_{n+2}-s_{n+1}\right) \subseteq E_{n}$ for every sufficiently large $n$.
Proof. For sufficiently large $n$, all elements of $\left(\sigma_{n}, s_{n+2}\right)$ are clean. Clearly, the elements of $\left(\sigma_{n}, s_{n+2}\right)$ can be in neither $O_{n}$ nor $O \backslash O_{n+1}$. So, $\left(\sigma_{n}, s_{n+2}\right) \subseteq O_{n+1} \backslash O_{n}=s_{n+1}+E_{n}$, which yields the claim.

Now, Claim 1 together with inequalities (3) implies that, for all sufficiently large $n$,

$$
E \supseteq E_{n} \supseteq\left(\sigma_{n}-s_{n+1}, s_{n+2}-s_{n+1}\right) \supseteq\left(2 n m, 2^{n-1}-(n+2) m\right) .
$$

This easily yields that $\mathbb{Z}_{\geqslant 0} \backslash E$ is also finite. Since $\mathbb{Z}_{\geqslant 0} \backslash O$ is also finite, by Property 1 , there exists a positive integer $N$ such that every integer $n>N$ has exactly one even and one odd representation.

Step 3. We investigate the structures of $E_{n}$ and $O_{n}$.
Suppose that $z \in E_{2 n}$. Since $z$ can be represented as an even sum using $\left\{s_{1}, s_{2}, \ldots, s_{2 n}\right\}$, so can its complement $\sigma_{2 n}-z$. Thus, we get $E_{2 n}=\sigma_{2 n}-E_{2 n}$. Similarly, we have

$$
\begin{equation*}
E_{2 n}=\sigma_{2 n}-E_{2 n}, \quad O_{2 n}=\sigma_{2 n}-O_{2 n}, \quad E_{2 n+1}=\sigma_{2 n+1}-O_{2 n+1}, \quad O_{2 n+1}=\sigma_{2 n+1}-E_{2 n+1} . \tag{4}
\end{equation*}
$$

Claim 2. For every sufficiently large $n$, we have

$$
\left[0, \sigma_{n}\right] \supseteq O_{n} \supseteq\left(N, \sigma_{n}-N\right) \quad \text { and } \quad\left[0, \sigma_{n}\right] \supseteq E_{n} \supseteq\left(N, \sigma_{n}-N\right)
$$

Proof. Clearly $O_{n}, E_{n} \subseteq\left[0, \sigma_{n}\right]$ for every positive integer $n$. We now prove $O_{n}, E_{n} \supseteq\left(N, \sigma_{n}-N\right)$. Taking $n$ sufficiently large, we may assume that $s_{n+1} \geqslant 2^{n-1}+1-n m>\frac{1}{2}\left(2^{n-1}-1+n m\right) \geqslant \sigma_{n} / 2$. Therefore, the odd representation of every element of ( $N, \sigma_{n} / 2$ ] cannot contain a term larger than $s_{n}$. Thus, $\left(N, \sigma_{n} / 2\right] \subseteq O_{n}$. Similarly, since $s_{n+1}+s_{1}>\sigma_{n} / 2$, we also have $\left(N, \sigma_{n} / 2\right] \subseteq E_{n}$. Equations (4) then yield that, for sufficiently large $n$, the interval $\left(N, \sigma_{n}-N\right)$ is a subset of both $O_{n}$ and $E_{n}$, as desired.

Step 4. We obtain a final contradiction.
Notice that $0 \in \mathbb{Z}_{\geqslant 0} \backslash O$ and $1 \in \mathbb{Z}_{\geqslant 0} \backslash E$. Therefore, the sets $\mathbb{Z}_{\geqslant 0} \backslash O$ and $\mathbb{Z}_{\geqslant 0} \backslash E$ are nonempty. Denote $o=\max \left(\mathbb{Z}_{\geqslant 0} \backslash O\right)$ and $e=\max \left(\mathbb{Z}_{\geqslant 0} \backslash E\right)$. Observe also that $e, o \leqslant N$.

Taking $k$ sufficiently large, we may assume that $\sigma_{2 k}>2 N$ and that Claim 2 holds for all $n \geqslant 2 k$. Due to (4) and Claim 2, we have that $\sigma_{2 k}-e$ is the minimal number greater than $N$ which is not in $E_{2 k}$, i.e., $\sigma_{2 k}-e=s_{2 k+1}+s_{1}$. Similarly,

$$
\sigma_{2 k}-o=s_{2 k+1}, \quad \sigma_{2 k+1}-e=s_{2 k+2}, \quad \text { and } \quad \sigma_{2 k+1}-o=s_{2 k+2}+s_{1}
$$

Therefore, we have

$$
\begin{aligned}
s_{1} & =\left(s_{2 k+1}+s_{1}\right)-s_{2 k+1}=\left(\sigma_{2 k}-e\right)-\left(\sigma_{2 k}-o\right)=o-e \\
& =\left(\sigma_{2 k+1}-e\right)-\left(\sigma_{2 k+1}-o\right)=s_{2 k+2}-\left(s_{2 k+2}+s_{1}\right)=-s_{1},
\end{aligned}
$$

which is impossible since $s_{1}>0$.

C7. In a company of people some pairs are enemies. A group of people is called unsociable if the number of members in the group is odd and at least 3 , and it is possible to arrange all its members around a round table so that every two neighbors are enemies. Given that there are at most 2015 unsociable groups, prove that it is possible to partition the company into 11 parts so that no two enemies are in the same part.
(Russia)
Solution 1. Let $G=(V, E)$ be a graph where $V$ is the set of people in the company and $E$ is the set of the enemy pairs - the edges of the graph. In this language, partitioning into 11 disjoint enemy-free subsets means properly coloring the vertices of this graph with 11 colors.

We will prove the following more general statement.
Claim. Let $G$ be a graph with chromatic number $k \geqslant 3$. Then $G$ contains at least $2^{k-1}-k$ unsociable groups.

Recall that the chromatic number of $G$ is the least $k$ such that a proper coloring

$$
\begin{equation*}
V=V_{1} \sqcup \cdots \sqcup V_{k} \tag{1}
\end{equation*}
$$

exists. In view of $2^{11}-12>2015$, the claim implies the problem statement.
Let $G$ be a graph with chromatic number $k$. We say that a proper coloring (1) of $G$ is leximinimal, if the $k$-tuple $\left(\left|V_{1}\right|,\left|V_{2}\right|, \ldots,\left|V_{k}\right|\right)$ is lexicographically minimal; in other words, the following conditions are satisfied: the number $n_{1}=\left|V_{1}\right|$ is minimal; the number $n_{2}=\left|V_{2}\right|$ is minimal, subject to the previously chosen value of $n_{1} ; \ldots$; the number $n_{k-1}=\left|V_{k-1}\right|$ is minimal, subject to the previously chosen values of $n_{1}, \ldots, n_{k-2}$.

The following lemma is the core of the proof.
Lemma 1. Suppose that $G=(V, E)$ is a graph with odd chromatic number $k \geqslant 3$, and let (1) be one of its leximinimal colorings. Then $G$ contains an odd cycle which visits all color classes $V_{1}, V_{2}, \ldots, V_{k}$.
Proof of Lemma 1. Let us call a cycle colorful if it visits all color classes.
Due to the definition of the chromatic number, $V_{1}$ is nonempty. Choose an arbitrary vertex $v \in V_{1}$. We construct a colorful odd cycle that has only one vertex in $V_{1}$, and this vertex is $v$.

We draw a subgraph of $G$ as follows. Place $v$ in the center, and arrange the sets $V_{2}, V_{3}, \ldots, V_{k}$ in counterclockwise circular order around it. For convenience, let $V_{k+1}=V_{2}$. We will draw arrows to add direction to some edges of $G$, and mark the vertices these arrows point to. First we draw arrows from $v$ to all its neighbors in $V_{2}$, and mark all those neighbors. If some vertex $u \in V_{i}$ with $i \in\{2,3, \ldots, k\}$ is already marked, we draw arrows from $u$ to all its neighbors in $V_{i+1}$ which are not marked yet, and we mark all of them. We proceed doing this as long as it is possible. The process of marking is exemplified in Figure 1.

Notice that by the rules of our process, in the final state, marked vertices in $V_{i}$ cannot have unmarked neighbors in $V_{i+1}$. Moreover, $v$ is connected to all marked vertices by directed paths.

Now move each marked vertex to the next color class in circular order (see an example in Figure 3). In view of the arguments above, the obtained coloring $V_{1} \sqcup W_{2} \sqcup \cdots \sqcup W_{k}$ is proper. Notice that $v$ has a neighbor $w \in W_{2}$, because otherwise

$$
\left(V_{1} \backslash\{v\}\right) \sqcup\left(W_{2} \cup\{v\}\right) \sqcup W_{3} \sqcup \cdots \sqcup W_{k}
$$

would be a proper coloring lexicographically smaller than (1). If $w$ was unmarked, i.e., $w$ was an element of $V_{2}$, then it would be marked at the beginning of the process and thus moved to $V_{3}$, which did not happen. Therefore, $w$ is marked and $w \in V_{k}$.


Figure 1
Since $w$ is marked, there exists a directed path from $v$ to $w$. This path moves through the sets $V_{2}, \ldots, V_{k}$ in circular order, so the number of edges in it is divisible by $k-1$ and thus even. Closing this path by the edge $w \rightarrow v$, we get a colorful odd cycle, as required.

Proof of the claim. Let us choose a leximinimal coloring (1) of $G$. For every set $C \subseteq\{1,2, \ldots, k\}$ such that $|C|$ is odd and greater than 1 , we will provide an odd cycle visiting exactly those color classes whose indices are listed in the set $C$. This property ensures that we have different cycles for different choices of $C$, and it proves the claim because there are $2^{k-1}-k$ choices for the set $C$.

Let $V_{C}=\bigcup_{c \in C} V_{c}$, and let $G_{C}$ be the induced subgraph of $G$ on the vertex set $V_{C}$. We also have the induced coloring of $V_{C}$ with $|C|$ colors; this coloring is of course proper. Notice further that the induced coloring is leximinimal: if we had a lexicographically smaller coloring $\left(W_{c}\right)_{c \in C}$ of $G_{C}$, then these classes, together the original color classes $V_{i}$ for $i \notin C$, would provide a proper coloring which is lexicographically smaller than (1). Hence Lemma 1, applied to the subgraph $G_{C}$ and its leximinimal coloring $\left(V_{c}\right)_{c \in C}$, provides an odd cycle that visits exactly those color classes that are listed in the set $C$.

Solution 2. We provide a different proof of the claim from the previous solution.
We say that a graph is critical if deleting any vertex from the graph decreases the graph's chromatic number. Obviously every graph contains a critical induced subgraph with the same chromatic number.
Lemma 2. Suppose that $G=(V, E)$ is a critical graph with chromatic number $k \geqslant 3$. Then every vertex $v$ of $G$ is contained in at least $2^{k-2}-1$ unsociable groups.
Proof. For every set $X \subseteq V$, denote by $n(X)$ the number of neighbors of $v$ in the set $X$.
Since $G$ is critical, there exists a proper coloring of $G \backslash\{v\}$ with $k-1$ colors, so there exists a proper coloring $V=V_{1} \sqcup V_{2} \sqcup \cdots \sqcup V_{k}$ of $G$ such that $V_{1}=\{v\}$. Among such colorings, take one for which the sequence $\left(n\left(V_{2}\right), n\left(V_{3}\right), \ldots, n\left(V_{k}\right)\right)$ is lexicographically minimal. Clearly, $n\left(V_{i}\right)>0$ for every $i=2,3, \ldots, k$; otherwise $V_{2} \sqcup \ldots \sqcup V_{i-1} \sqcup\left(V_{i} \cup V_{1}\right) \sqcup V_{i+1} \sqcup \ldots V_{k}$ would be a proper coloring of $G$ with $k-1$ colors.

We claim that for every $C \subseteq\{2,3, \ldots, k\}$ with $|C| \geqslant 2$ being even, $G$ contains an unsociable group so that the set of its members' colors is precisely $C \cup\{1\}$. Since the number of such sets $C$ is $2^{k-2}-1$, this proves the lemma. Denote the elements of $C$ by $c_{1}, \ldots, c_{2 \ell}$ in increasing order. For brevity, let $U_{i}=V_{c_{i}}$. Denote by $N_{i}$ the set of neighbors of $v$ in $U_{i}$.

We show that for every $i=1, \ldots, 2 \ell-1$ and $x \in N_{i}$, the subgraph induced by $U_{i} \cup U_{i+1}$ contains a path that connects $x$ with another point in $N_{i+1}$. For the sake of contradiction, suppose that no such path exists. Let $S$ be the set of vertices that lie in the connected component of $x$ in the subgraph induced by $U_{i} \cup U_{i+1}$, and let $P=U_{i} \cap S$, and $Q=U_{i+1} \cap S$ (see Figure 3). Since $x$ is separated from $N_{i+1}$, the sets $Q$ and $N_{i+1}$ are disjoint. So, if we re-color $G$ by replacing $U_{i}$ and $U_{i+1}$ by $\left(U_{i} \cup Q\right) \backslash P$ and $\left(U_{i+1} \cup P\right) \backslash Q$, respectively, we obtain a proper coloring such that $n\left(U_{i}\right)=n\left(V_{c_{i}}\right)$ is decreased and only $n\left(U_{i+1}\right)=n\left(V_{c_{i+1}}\right)$ is increased. That contradicts the lexicographical minimality of $\left(n\left(V_{2}\right), n\left(V_{3}\right), \ldots, n\left(V_{k}\right)\right)$.


Figure 3
Next, we build a path through $U_{1}, U_{2}, \ldots, U_{2 \ell}$ as follows. Let the starting point of the path be an arbitrary vertex $v_{1}$ in the set $N_{1}$. For $i \leqslant 2 \ell-1$, if the vertex $v_{i} \in N_{i}$ is already defined, connect $v_{i}$ to some vertex in $N_{i+1}$ in the subgraph induced by $U_{i} \cup U_{i+1}$, and add these edges to the path. Denote the new endpoint of the path by $v_{i+1}$; by the construction we have $v_{i+1} \in N_{i+1}$ again, so the process can be continued. At the end we have a path that starts at $v_{1} \in N_{1}$ and ends at some $v_{2 \ell} \in N_{2 \ell}$. Moreover, all edges in this path connect vertices in neighboring classes: if a vertex of the path lies in $U_{i}$, then the next vertex lies in $U_{i+1}$ or $U_{i-1}$. Notice that the path is not necessary simple, so take a minimal subpath of it. The minimal subpath is simple and connects the same endpoints $v_{1}$ and $v_{2 \ell}$. The property that every edge steps to a neighboring color class (i.e., from $U_{i}$ to $U_{i+1}$ or $U_{i-1}$ ) is preserved. So the resulting path also visits all of $U_{1}, \ldots, U_{2 \ell}$, and its length must be odd. Closing the path with the edges $v v_{1}$ and $v_{2 \ell} v$ we obtain the desired odd cycle (see Figure 4).


Figure 4
Now we prove the claim by induction on $k \geqslant 3$. The base case $k=3$ holds by applying Lemma 2 to a critical subgraph. For the induction step, let $G_{0}$ be a critical $k$-chromatic subgraph of $G$, and let $v$ be an arbitrary vertex of $G_{0}$. By Lemma $2, G_{0}$ has at least $2^{k-2}-1$ unsociable groups containing $v$. On the other hand, the graph $G_{0} \backslash\{v\}$ has chromatic number $k-1$, so it contains at least $2^{k-2}-(k-1)$ unsociable groups by the induction hypothesis. Altogether, this gives $2^{k-2}-1+2^{k-2}-(k-1)=2^{k-1}-k$ distinct unsociable groups in $G_{0}$ (and thus in $G$ ).

Comment 1. The claim we proved is sharp. The complete graph with $k$ vertices has chromatic number $k$ and contains exactly $2^{k-1}-k$ unsociable groups.

Comment 2. The proof of Lemma 2 works for odd values of $|C| \geqslant 3$ as well. Hence, the second solution shows the analogous statement that the number of even sized unsociable groups is at least $2^{k}-1-\binom{k}{2}$.

## Geometry

G1. Let $A B C$ be an acute triangle with orthocenter $H$. Let $G$ be the point such that the quadrilateral $A B G H$ is a parallelogram. Let $I$ be the point on the line $G H$ such that $A C$ bisects $H I$. Suppose that the line $A C$ intersects the circumcircle of the triangle $G C I$ at $C$ and $J$. Prove that $I J=A H$.
(Australia)
Solution 1. Since $H G \| A B$ and $B G \| A H$, we have $B G \perp B C$ and $C H \perp G H$. Therefore, the quadrilateral $B G C H$ is cyclic. Since $H$ is the orthocenter of the triangle $A B C$, we have $\angle H A C=90^{\circ}-\angle A C B=\angle C B H$. Using that $B G C H$ and $C G J I$ are cyclic quadrilaterals, we get

$$
\angle C J I=\angle C G H=\angle C B H=\angle H A C .
$$

Let $M$ be the intersection of $A C$ and $G H$, and let $D \neq A$ be the point on the line $A C$ such that $A H=H D$. Then $\angle M J I=\angle H A C=\angle M D H$.

Since $\angle M J I=\angle M D H, \angle I M J=\angle H M D$, and $I M=M H$, the triangles $I M J$ and $H M D$ are congruent, and thus $I J=H D=A H$.


Comment. Instead of introducing the point $D$, one can complete the solution by using the law of sines in the triangles $I J M$ and $A M H$, yielding

$$
\frac{I J}{I M}=\frac{\sin \angle I M J}{\sin \angle M J I}=\frac{\sin \angle A M H}{\sin \angle H A M}=\frac{A H}{M H}=\frac{A H}{I M} .
$$

Solution 2. Obtain $\angle C G H=\angle H A C$ as in the previous solution. In the parallelogram $A B G H$ we have $\angle B A H=\angle H G B$. It follows that

$$
\angle H M C=\angle B A C=\angle B A H+\angle H A C=\angle H G B+\angle C G H=\angle C G B .
$$

So the right triangles $C M H$ and $C G B$ are similar. Also, in the circumcircle of triangle $G C I$ we have similar triangles $M I J$ and $M C G$. Therefore,

$$
\frac{I J}{C G}=\frac{M I}{M C}=\frac{M H}{M C}=\frac{G B}{G C}=\frac{A H}{C G} .
$$

Hence $I J=A H$.

G2. Let $A B C$ be a triangle inscribed into a circle $\Omega$ with center $O$. A circle $\Gamma$ with center $A$ meets the side $B C$ at points $D$ and $E$ such that $D$ lies between $B$ and $E$. Moreover, let $F$ and $G$ be the common points of $\Gamma$ and $\Omega$. We assume that $F$ lies on the arc $A B$ of $\Omega$ not containing $C$, and $G$ lies on the arc $A C$ of $\Omega$ not containing $B$. The circumcircles of the triangles $B D F$ and $C E G$ meet the sides $A B$ and $A C$ again at $K$ and $L$, respectively. Suppose that the lines $F K$ and $G L$ are distinct and intersect at $X$. Prove that the points $A, X$, and $O$ are collinear.
(Greece)
Solution 1. It suffices to prove that the lines $F K$ and $G L$ are symmetric about $A O$. Now the segments $A F$ and $A G$, being chords of $\Omega$ with the same length, are clearly symmetric with respect to $A O$. Hence it is enough to show

$$
\begin{equation*}
\angle K F A=\angle A G L \tag{1}
\end{equation*}
$$

Let us denote the circumcircles of $B D F$ and $C E G$ by $\omega_{B}$ and $\omega_{C}$, respectively. To prove (1), we start from

$$
\angle K F A=\angle D F G+\angle G F A-\angle D F K .
$$

In view of the circles $\omega_{B}, \Gamma$, and $\Omega$, this may be rewritten as

$$
\angle K F A=\angle C E G+\angle G B A-\angle D B K=\angle C E G-\angle C B G .
$$

Due to the circles $\omega_{C}$ and $\Omega$, we obtain $\angle K F A=\angle C L G-\angle C A G=\angle A G L$. Thereby the problem is solved.


Figure 1

Solution 2. Again, we denote the circumcircle of $B D K F$ by $\omega_{B}$. In addition, we set $\alpha=$ $\angle B A C, \varphi=\angle A B F$, and $\psi=\angle E D A=\angle A E D$ (see Figure 2). Notice that $A F=A G$ entails $\varphi=\angle G C A$, so all three of $\alpha, \varphi$, and $\psi$ respect the "symmetry" between $B$ and $C$ of our configuration. Again, we reduce our task to proving (1).

This time, we start from

$$
2 \angle K F A=2(\angle D F A-\angle D F K) .
$$

Since the triangle $A F D$ is isosceles, we have

$$
\angle D F A=\angle A D F=\angle E D F-\psi=\angle B F D+\angle E B F-\psi
$$

Moreover, because of the circle $\omega_{B}$ we have $\angle D F K=\angle C B A$. Altogether, this yields

$$
2 \angle K F A=\angle D F A+(\angle B F D+\angle E B F-\psi)-2 \angle C B A,
$$

which simplifies to

$$
2 \angle K F A=\angle B F A+\varphi-\psi-\angle C B A .
$$

Now the quadrilateral $A F B C$ is cyclic, so this entails $2 \angle K F A=\alpha+\varphi-\psi$.
Due to the "symmetry" between $B$ and $C$ alluded to above, this argument also shows that $2 \angle A G L=\alpha+\varphi-\psi$. This concludes the proof of (1).


Figure 2

Comment 1. As the first solution shows, the assumption that $A$ be the center of $\Gamma$ may be weakened to the following one: The center of $\Gamma$ lies on the line $O A$. The second solution may be modified to yield the same result.

Comment 2. It might be interesting to remark that $\angle G D K=90^{\circ}$. To prove this, let $G^{\prime}$ denote the point on $\Gamma$ diametrically opposite to $G$. Because of $\angle K D F=\angle K B F=\angle A G F=\angle G^{\prime} D F$, the points $D, K$, and $G^{\prime}$ are collinear, which leads to the desired result. Notice that due to symmetry we also have $\angle L E F=90^{\circ}$.

Moreover, a standard argument shows that the triangles $A G L$ and $B G E$ are similar. By symmetry again, also the triangles $A F K$ and $C D F$ are similar.

There are several ways to derive a solution from these facts. For instance, one may argue that

$$
\begin{aligned}
\angle K F A & =\angle B F A-\angle B F K=\angle B F A-\angle E D G^{\prime}=\left(180^{\circ}-\angle A G B\right)-\left(180^{\circ}-\angle G^{\prime} G E\right) \\
& =\angle A G E-\angle A G B=\angle B G E=\angle A G L .
\end{aligned}
$$

Comment 3. The original proposal did not contain the point $X$ in the assumption and asked instead to prove that the lines $F K, G L$, and $A O$ are concurrent. This differs from the version given above only insofar as it also requires to show that these lines cannot be parallel. The Problem Selection Committee removed this part from the problem intending to make it thus more suitable for the Olympiad.

For the sake of completeness, we would still like to sketch one possibility for proving $F K \nVdash A O$ here. As the points $K$ and $O$ lie in the angular region $\angle F A G$, it suffices to check $\angle K F A+\angle F A O<180^{\circ}$. Multiplying by 2 and making use of the formulae from the second solution, we see that this is equivalent to $(\alpha+\varphi-\psi)+\left(180^{\circ}-2 \varphi\right)<360^{\circ}$, which in turn is an easy consequence of $\alpha<180^{\circ}$.

G3. Let $A B C$ be a triangle with $\angle C=90^{\circ}$, and let $H$ be the foot of the altitude from $C$. A point $D$ is chosen inside the triangle $C B H$ so that $C H$ bisects $A D$. Let $P$ be the intersection point of the lines $B D$ and $C H$. Let $\omega$ be the semicircle with diameter $B D$ that meets the segment $C B$ at an interior point. A line through $P$ is tangent to $\omega$ at $Q$. Prove that the lines $C Q$ and $A D$ meet on $\omega$.
(Georgia)
Solution 1. Let $K$ be the projection of $D$ onto $A B$; then $A H=H K$ (see Figure 1). Since $P H \| D K$, we have

$$
\begin{equation*}
\frac{P D}{P B}=\frac{H K}{H B}=\frac{A H}{H B} \tag{1}
\end{equation*}
$$

Let $L$ be the projection of $Q$ onto $D B$. Since $P Q$ is tangent to $\omega$ and $\angle D Q B=\angle B L Q=$ $90^{\circ}$, we have $\angle P Q D=\angle Q B P=\angle D Q L$. Therefore, $Q D$ and $Q B$ are respectively the internal and the external bisectors of $\angle P Q L$. By the angle bisector theorem, we obtain

$$
\begin{equation*}
\frac{P D}{D L}=\frac{P Q}{Q L}=\frac{P B}{B L} . \tag{2}
\end{equation*}
$$

The relations (1) and (2) yield $\frac{A H}{H B}=\frac{P D}{P B}=\frac{D L}{L B}$. So, the spiral similarity $\tau$ centered at $B$ and sending $A$ to $D$ maps $H$ to $L$. Moreover, $\tau$ sends the semicircle with diameter $A B$ passing through $C$ to $\omega$. Due to $C H \perp A B$ and $Q L \perp D B$, it follows that $\tau(C)=Q$.

Hence, the triangles $A B D$ and $C B Q$ are similar, so $\angle A D B=\angle C Q B$. This means that the lines $A D$ and $C Q$ meet at some point $T$, and this point satisfies $\angle B D T=\angle B Q T$. Therefore, $T$ lies on $\omega$, as needed.


Figure 1


Figure 2

Comment 1. Since $\angle B A D=\angle B C Q$, the point $T$ lies also on the circumcircle of the triangle $A B C$.
Solution 2. Let $\Gamma$ be the circumcircle of $A B C$, and let $A D$ meet $\omega$ at $T$. Then $\angle A T B=$ $\angle A C B=90^{\circ}$, so $T$ lies on $\Gamma$ as well. As in the previous solution, let $K$ be the projection of $D$ onto $A B$; then $A H=H K$ (see Figure 2).

Our goal now is to prove that the points $C, Q$, and $T$ are collinear. Let $C T$ meet $\omega$ again at $Q^{\prime}$. Then, it suffices to show that $P Q^{\prime}$ is tangent to $\omega$, or that $\angle P Q^{\prime} D=\angle Q^{\prime} B D$.

Since the quadrilateral $B D Q^{\prime} T$ is cyclic and the triangles $A H C$ and $K H C$ are congruent, we have $\angle Q^{\prime} B D=\angle Q^{\prime} T D=\angle C T A=\angle C B A=\angle A C H=\angle H C K$. Hence, the right triangles $C H K$ and $B Q^{\prime} D$ are similar. This implies that $\frac{H K}{C K}=\frac{Q^{\prime} D}{B D}$, and thus $H K \cdot B D=C K \cdot Q^{\prime} D$. Notice that $P H \| D K$; therefore, we have $\frac{P D}{B D}=\frac{H K}{B K}$, and so $P D \cdot B K=H K \cdot B D$. Consequently, $P D \cdot B K=H K \cdot B D=C K \cdot Q^{\prime} D$, which yields $\frac{P D}{Q^{\prime} D}=\frac{C K}{B K}$.

Since $\angle C K A=\angle K A C=\angle B D Q^{\prime}$, the triangles $C K B$ and $P D Q^{\prime}$ are similar, so $\angle P Q^{\prime} D=$ $\angle C B A=\angle Q^{\prime} B D$, as required.

Comment 2. There exist several other ways to prove that $P Q^{\prime}$ is tangent to $\omega$. For instance, one may compute $\frac{P D}{P B}$ and $\frac{P Q^{\prime}}{P B}$ in terms of $A H$ and $H B$ to verify that $P Q^{\prime 2}=P D \cdot P B$, concluding that $P Q^{\prime}$ is tangent to $\omega$.

Another possible approach is the following. As in Solution 2, we introduce the points $T$ and $Q^{\prime}$ and mention that the triangles $A B C$ and $D B Q^{\prime}$ are similar (see Figure 3).

Let $M$ be the midpoint of $A D$, and let $L$ be the projection of $Q^{\prime}$ onto $A B$. Construct $E$ on the line $A B$ so that $E P$ is parallel to $A D$. Projecting from $P$, we get $(A, B ; H, E)=(A, D ; M, \infty)=-1$.

Since $\frac{E A}{A B}=\frac{P D}{D B}$, the point $P$ is the image of $E$ under the similarity transform mapping $A B C$ to $D B Q^{\prime}$. Therefore, we have $(D, B ; L, P)=(A, B ; H, E)=-1$, which means that $Q^{\prime} D$ and $Q^{\prime} B$ are respectively the internal and the external bisectors of $\angle P Q^{\prime} L$. This implies that $P Q^{\prime}$ is tangent to $\omega$, as required.


Figure 3
Solution 3. Introduce the points $T$ and $Q^{\prime}$ as in the previous solution. Note that $T$ lies on the circumcircle of $A B C$. Here we present yet another proof that $P Q^{\prime}$ is tangent to $\omega$.

Let $\Omega$ be the circle completing the semicircle $\omega$. Construct a point $F$ symmetric to $C$ with respect to $A B$. Let $S \neq T$ be the second intersection point of $F T$ and $\Omega$ (see Figure 4).


Figure 4
Since $A C=A F$, we have $\angle D K C=\angle H C K=\angle C B A=\angle C T A=\angle D T S=180^{\circ}-$ $\angle S K D$. Thus, the points $C, K$, and $S$ are collinear. Notice also that $\angle Q^{\prime} K D=\angle Q^{\prime} T D=$ $\angle H C K=\angle K F H=180^{\circ}-\angle D K F$. This implies that the points $F, K$, and $Q^{\prime}$ are collinear.

Applying Pascal's theorem to the degenerate hexagon $K Q^{\prime} Q^{\prime} T S S$, we get that the tangents to $\Omega$ passing through $Q^{\prime}$ and $S$ intersect on $C F$. The relation $\angle Q^{\prime} T D=\angle D T S$ yields that $Q^{\prime}$ and $S$ are symmetric with respect to $B D$. Therefore, the two tangents also intersect on $B D$. Thus, the two tangents pass through $P$. Hence, $P Q^{\prime}$ is tangent to $\omega$, as needed.

G4. Let $A B C$ be an acute triangle, and let $M$ be the midpoint of $A C$. A circle $\omega$ passing through $B$ and $M$ meets the sides $A B$ and $B C$ again at $P$ and $Q$, respectively. Let $T$ be the point such that the quadrilateral $B P T Q$ is a parallelogram. Suppose that $T$ lies on the circumcircle of the triangle $A B C$. Determine all possible values of $B T / B M$.
(Russia)
Answer. $\sqrt{2}$.
Solution 1. Let $S$ be the center of the parallelogram $B P T Q$, and let $B^{\prime} \neq B$ be the point on the ray $B M$ such that $B M=M B^{\prime}$ (see Figure 1). It follows that $A B C B^{\prime}$ is a parallelogram. Then, $\angle A B B^{\prime}=\angle P Q M$ and $\angle B B^{\prime} A=\angle B^{\prime} B C=\angle M P Q$, and so the triangles $A B B^{\prime}$ and $M Q P$ are similar. It follows that $A M$ and $M S$ are corresponding medians in these triangles. Hence,

$$
\begin{equation*}
\angle S M P=\angle B^{\prime} A M=\angle B C A=\angle B T A . \tag{1}
\end{equation*}
$$

Since $\angle A C T=\angle P B T$ and $\angle T A C=\angle T B C=\angle B T P$, the triangles $T C A$ and $P B T$ are similar. Again, as $T M$ and $P S$ are corresponding medians in these triangles, we have

$$
\begin{equation*}
\angle M T A=\angle T P S=\angle B Q P=\angle B M P \tag{2}
\end{equation*}
$$

Now we deal separately with two cases.
Case 1. $S$ does not lie on $B M$. Since the configuration is symmetric between $A$ and $C$, we may assume that $S$ and $A$ lie on the same side with respect to the line $B M$.

Applying (1) and (2), we get

$$
\angle B M S=\angle B M P-\angle S M P=\angle M T A-\angle B T A=\angle M T B
$$

and so the triangles $B S M$ and $B M T$ are similar. We now have $B M^{2}=B S \cdot B T=B T^{2} / 2$, so $B T=\sqrt{2} B M$.

Case 2. $S$ lies on $B M$. It follows from (2) that $\angle B C A=\angle M T A=\angle B Q P=\angle B M P$ (see Figure 2). Thus, $P Q \| A C$ and $P M \| A T$. Hence, $B S / B M=B P / B A=B M / B T$, so $B T^{2}=2 B M^{2}$ and $B T=\sqrt{2} B M$.


Figure 1


Figure 2

Comment 1. Here is another way to show that the triangles $B S M$ and $B M T$ are similar. Denote by $\Omega$ the circumcircle of the triangle $A B C$. Let $R$ be the second point of intersection of $\omega$ and $\Omega$, and let $\tau$ be the spiral similarity centered at $R$ mapping $\omega$ to $\Omega$. Then, one may show that $\tau$ maps each point $X$ on $\omega$ to a point $Y$ on $\Omega$ such that $B, X$, and $Y$ are collinear (see Figure 3). If we let $K$ and $L$ be the second points of intersection of $B M$ with $\Omega$ and of $B T$ with $\omega$, respectively, then it follows that the triangle $M K T$ is the image of $S M L$ under $\tau$. We now obtain $\angle B S M=\angle T M B$, which implies the desired result.


Figure 3


Figure 4

Solution 2. Again, we denote by $\Omega$ the circumcircle of the triangle $A B C$.
Choose the points $X$ and $Y$ on the rays $B A$ and $B C$ respectively, so that $\angle M X B=\angle M B C$ and $\angle B Y M=\angle A B M$ (see Figure 4). Then the triangles $B M X$ and $Y M B$ are similar. Since $\angle X P M=\angle B Q M$, the points $P$ and $Q$ correspond to each other in these triangles. So, if $\overrightarrow{B P}=\mu \cdot \overrightarrow{B X}$, then $\overrightarrow{B Q}=(1-\mu) \cdot \overrightarrow{B Y}$. Thus

$$
\overrightarrow{B T}=\overrightarrow{B P}+\overrightarrow{B Q}=\overrightarrow{B Y}+\mu \cdot(\overrightarrow{B X}-\overrightarrow{B Y})=\overrightarrow{B Y}+\mu \cdot \overrightarrow{Y X},
$$

which means that $T$ lies on the line $X Y$.
Let $B^{\prime} \neq B$ be the point on the ray $B M$ such that $B M=M B^{\prime}$. Then $\angle M B^{\prime} A=$ $\angle M B C=\angle M X B$ and $\angle C B^{\prime} M=\angle A B M=\angle B Y M$. This means that the triangles $B M X$, $B A B^{\prime}, Y M B$, and $B^{\prime} C B$ are all similar; hence $B A \cdot B X=B M \cdot B B^{\prime}=B C \cdot B Y$. Thus there exists an inversion centered at $B$ which swaps $A$ with $X, M$ with $B^{\prime}$, and $C$ with $Y$. This inversion then swaps $\Omega$ with the line $X Y$, and hence it preserves $T$. Therefore, we have $B T^{2}=B M \cdot B B^{\prime}=2 B M^{2}$, and $B T=\sqrt{2} B M$.

Solution 3. We begin with the following lemma.
Lemma. Let $A B C T$ be a cyclic quadrilateral. Let $P$ and $Q$ be points on the sides $B A$ and $B C$ respectively, such that $B P T Q$ is a parallelogram. Then $B P \cdot B A+B Q \cdot B C=B T^{2}$.
Proof. Let the circumcircle of the triangle $Q T C$ meet the line $B T$ again at $J$ (see Figure 5). The power of $B$ with respect to this circle yields

$$
\begin{equation*}
B Q \cdot B C=B J \cdot B T \tag{3}
\end{equation*}
$$

We also have $\angle T J Q=180^{\circ}-\angle Q C T=\angle T A B$ and $\angle Q T J=\angle A B T$, and so the triangles $T J Q$ and $B A T$ are similar. We now have $T J / T Q=B A / B T$. Therefore,

$$
\begin{equation*}
T J \cdot B T=T Q \cdot B A=B P \cdot B A \tag{4}
\end{equation*}
$$

Combining (3) and (4) now yields the desired result.
Let $X$ and $Y$ be the midpoints of $B A$ and $B C$ respectively (see Figure 6). Applying the lemma to the cyclic quadrilaterals $P B Q M$ and $A B C T$, we obtain

$$
B X \cdot B P+B Y \cdot B Q=B M^{2}
$$

and

$$
B P \cdot B A+B Q \cdot B C=B T^{2}
$$

Since $B A=2 B X$ and $B C=2 B Y$, we have $B T^{2}=2 B M^{2}$, and so $B T=\sqrt{2} B M$.


Figure 5


Figure 6

Comment 2. Here we give another proof of the lemma using Ptolemy's theorem. We readily have

$$
T C \cdot B A+T A \cdot B C=A C \cdot B T .
$$

The lemma now follows from

$$
\frac{B P}{T C}=\frac{B Q}{T A}=\frac{B T}{A C}=\frac{\sin \angle B C T}{\sin \angle A B C} .
$$

G5. Let $A B C$ be a triangle with $C A \neq C B$. Let $D, F$, and $G$ be the midpoints of the sides $A B, A C$, and $B C$, respectively. A circle $\Gamma$ passing through $C$ and tangent to $A B$ at $D$ meets the segments $A F$ and $B G$ at $H$ and $I$, respectively. The points $H^{\prime}$ and $I^{\prime}$ are symmetric to $H$ and $I$ about $F$ and $G$, respectively. The line $H^{\prime} I^{\prime}$ meets $C D$ and $F G$ at $Q$ and $M$, respectively. The line $C M$ meets $\Gamma$ again at $P$. Prove that $C Q=Q P$.
(El Salvador)
Solution 1. We may assume that $C A>C B$. Observe that $H^{\prime}$ and $I^{\prime}$ lie inside the segments $C F$ and $C G$, respectively. Therefore, $M$ lies outside $\triangle A B C$ (see Figure 1).

Due to the powers of points $A$ and $B$ with respect to the circle $\Gamma$, we have

$$
C H^{\prime} \cdot C A=A H \cdot A C=A D^{2}=B D^{2}=B I \cdot B C=C I^{\prime} \cdot C B .
$$

Therefore, $C H^{\prime} \cdot C F=C I^{\prime} \cdot C G$. Hence, the quadrilateral $H^{\prime} I^{\prime} G F$ is cyclic, and so $\angle I^{\prime} H^{\prime} C=$ $\angle C G F$.

Let $D F$ and $D G$ meet $\Gamma$ again at $R$ and $S$, respectively. We claim that the points $R$ and $S$ lie on the line $H^{\prime} I^{\prime}$.

Observe that $F H^{\prime} \cdot F A=F H \cdot F C=F R \cdot F D$. Thus, the quadrilateral $A D H^{\prime} R$ is cyclic, and hence $\angle R H^{\prime} F=\angle F D A=\angle C G F=\angle I^{\prime} H^{\prime} C$. Therefore, the points $R, H^{\prime}$, and $I^{\prime}$ are collinear. Similarly, the points $S, H^{\prime}$, and $I^{\prime}$ are also collinear, and so all the points $R, H^{\prime}, Q, I^{\prime}, S$, and $M$ are all collinear.


Figure 1


Figure 2

Then, $\angle R S D=\angle R D A=\angle D F G$. Hence, the quadrilateral $R S G F$ is cyclic (see Figure 2). Therefore, $M H^{\prime} \cdot M I^{\prime}=M F \cdot M G=M R \cdot M S=M P \cdot M C$. Thus, the quadrilateral $C P I^{\prime} H^{\prime}$ is also cyclic. Let $\omega$ be its circumcircle.

Notice that $\angle H^{\prime} C Q=\angle S D C=\angle S R C$ and $\angle Q C I^{\prime}=\angle C D R=\angle C S R$. Hence, $\triangle C H^{\prime} Q \sim \triangle R C Q$ and $\triangle C I^{\prime} Q \sim \triangle S C Q$, and therefore $Q H^{\prime} \cdot Q R=Q C^{2}=Q I^{\prime} \cdot Q S$.

We apply the inversion with center $Q$ and radius $Q C$. Observe that the points $R, C$, and $S$ are mapped to $H^{\prime}, C$, and $I^{\prime}$, respectively. Therefore, the circumcircle $\Gamma$ of $\triangle R C S$ is mapped to the circumcircle $\omega$ of $\triangle H^{\prime} C I^{\prime}$. Since $P$ and $C$ belong to both circles and the point $C$ is preserved by the inversion, we have that $P$ is also mapped to itself. We then get $Q P^{2}=Q C^{2}$. Hence, $Q P=Q C$.

Comment 1. The problem statement still holds when $\Gamma$ intersects the sides $C A$ and $C B$ outside segments $A F$ and $B G$, respectively.

Solution 2. Let $X=H I \cap A B$, and let the tangent to $\Gamma$ at $C$ meet $A B$ at $Y$. Let $X C$ meet $\Gamma$ again at $X^{\prime}$ (see Figure 3). Projecting from $C, X$, and $C$ again, we have $(X, A ; D, B)=$ $\left(X^{\prime}, H ; D, I\right)=(C, I ; D, H)=(Y, B ; D, A)$. Since $A$ and $B$ are symmetric about $D$, it follows that $X$ and $Y$ are also symmetric about $D$.

Now, Menelaus' theorem applied to $\triangle A B C$ with the line $H I X$ yields

$$
1=\frac{C H}{H A} \cdot \frac{B I}{I C} \cdot \frac{A X}{X B}=\frac{A H^{\prime}}{H^{\prime} C} \cdot \frac{C I^{\prime}}{I^{\prime} B} \cdot \frac{B Y}{Y A} .
$$

By the converse of Menelaus' theorem applied to $\triangle A B C$ with points $H^{\prime}, I^{\prime}, Y$, we get that the points $H^{\prime}, I^{\prime}, Y$ are collinear.


Figure 3
Let $T$ be the midpoint of $C D$, and let $O$ be the center of $\Gamma$. Let $C M$ meet $T Y$ at $N$. To avoid confusion, we clean some superfluous details out of the picture (see Figure 4).

Let $V=M T \cap C Y$. Since $M T \| Y D$ and $D T=T C$, we get $C V=V Y$. Then Ceva's theorem applied to $\triangle C T Y$ with the point $M$ yields

$$
1=\frac{T Q}{Q C} \cdot \frac{C V}{V Y} \cdot \frac{Y N}{N T}=\frac{T Q}{Q C} \cdot \frac{Y N}{N T}
$$

Therefore, $\frac{T Q}{Q C}=\frac{T N}{N Y}$. So, $N Q \| C Y$, and thus $N Q \perp O C$.
Note that the points $O, N, T$, and $Y$ are collinear. Therefore, $C Q \perp O N$. So, $Q$ is the orthocenter of $\triangle O C N$, and hence $O Q \perp C P$. Thus, $Q$ lies on the perpendicular bisector of $C P$, and therefore $C Q=Q P$, as required.


Figure 4

Comment 2. The second part of Solution 2 provides a proof of the following more general statement, which does not involve a specific choice of $Q$ on $C D$.

Let YC and YD be two tangents to a circle $\Gamma$ with center $O$ (see Figure 4). Let $\ell$ be the midline of $\triangle Y C D$ parallel to $Y D$. Let $Q$ and $M$ be two points on $C D$ and $\ell$, respectively, such that the line $Q M$ passes through $Y$. Then $O Q \perp C M$.

G6. Let $A B C$ be an acute triangle with $A B>A C$, and let $\Gamma$ be its circumcircle. Let $H$, $M$, and $F$ be the orthocenter of the triangle, the midpoint of $B C$, and the foot of the altitude from $A$, respectively. Let $Q$ and $K$ be the two points on $\Gamma$ that satisfy $\angle A Q H=90^{\circ}$ and $\angle Q K H=90^{\circ}$. Prove that the circumcircles of the triangles $K Q H$ and $K F M$ are tangent to each other.
(Ukraine)
Solution 1. Let $A^{\prime}$ be the point diametrically opposite to $A$ on $\Gamma$. Since $\angle A Q A^{\prime}=90^{\circ}$ and $\angle A Q H=90^{\circ}$, the points $Q, H$, and $A^{\prime}$ are collinear. Similarly, if $Q^{\prime}$ denotes the point on $\Gamma$ diametrically opposite to $Q$, then $K, H$, and $Q^{\prime}$ are collinear. Let the line $A H F$ intersect $\Gamma$ again at $E$; it is known that $M$ is the midpoint of the segment $H A^{\prime}$ and that $F$ is the midpoint of $H E$. Let $J$ be the midpoint of $H Q^{\prime}$.

Consider any point $T$ such that $T K$ is tangent to the circle $K Q H$ at $K$ with $Q$ and $T$ lying on different sides of $K H$ (see Figure 1). Then $\angle H K T=\angle H Q K$ and we are to prove that $\angle M K T=\angle C F K$. Thus it remains to show that $\angle H Q K=\angle C F K+\angle H K M$. Due to $\angle H Q K=90^{\circ}-\angle Q^{\prime} H A^{\prime}$ and $\angle C F K=90^{\circ}-\angle K F A$, this means the same as $\angle Q^{\prime} H A^{\prime}=$ $\angle K F A-\angle H K M$. Now, since the triangles $K H E$ and $A H Q^{\prime}$ are similar with $F$ and $J$ being the midpoints of corresponding sides, we have $\angle K F A=\angle H J A$, and analogously one may obtain $\angle H K M=\angle J Q H$. Thereby our task is reduced to verifying

$$
\angle Q^{\prime} H A^{\prime}=\angle H J A-\angle J Q H .
$$



Figure 1


Figure 2

To avoid confusion, let us draw a new picture at this moment (see Figure 2). Owing to $\angle Q^{\prime} H A^{\prime}=\angle J Q H+\angle H J Q$ and $\angle H J A=\angle Q J A+\angle H J Q$, we just have to show that $2 \angle J Q H=\angle Q J A$. To this end, it suffices to remark that $A Q A^{\prime} Q^{\prime}$ is a rectangle and that $J$, being defined to be the midpoint of $H Q^{\prime}$, has to lie on the mid parallel of $Q A^{\prime}$ and $Q^{\prime} A$.

Solution 2. We define the points $A^{\prime}$ and $E$ and prove that the ray $M H$ passes through $Q$ in the same way as in the first solution. Notice that the points $A^{\prime}$ and $E$ can play analogous roles to the points $Q$ and $K$, respectively: point $A^{\prime}$ is the second intersection of the line $M H$ with $\Gamma$, and $E$ is the point on $\Gamma$ with the property $\angle H E A^{\prime}=90^{\circ}$ (see Figure 3).

In the circles $K Q H$ and $E A^{\prime} H$, the line segments $H Q$ and $H A^{\prime}$ are diameters, respectively; so, these circles have a common tangent $t$ at $H$, perpendicular to $M H$. Let $R$ be the radical center of the circles $A B C, K Q H$ and $E A^{\prime} H$. Their pairwise radical axes are the lines $Q K$, $A^{\prime} E$ and the line $t$; they all pass through $R$. Let $S$ be the midpoint of $H R$; by $\angle Q K H=$


Figure 3
$\angle H E A^{\prime}=90^{\circ}$, the quadrilateral $H E R K$ is cyclic and its circumcenter is $S$; hence we have $S K=S E=S H$. The line $B C$, being the perpendicular bisector of $H E$, passes through $S$.

The circle $H M F$ also is tangent to $t$ at $H$; from the power of $S$ with respect to the circle $H M F$ we have

$$
S M \cdot S F=S H^{2}=S K^{2} .
$$

So, the power of $S$ with respect to the circles $K Q H$ and $K F M$ is $S K^{2}$. Therefore, the line segment $S K$ is tangent to both circles at $K$.

G7. Let $A B C D$ be a convex quadrilateral, and let $P, Q, R$, and $S$ be points on the sides $A B, B C, C D$, and $D A$, respectively. Let the line segments $P R$ and $Q S$ meet at $O$. Suppose that each of the quadrilaterals $A P O S, B Q O P, C R O Q$, and $D S O R$ has an incircle. Prove that the lines $A C, P Q$, and $R S$ are either concurrent or parallel to each other.
(Bulgaria)
Solution 1. Denote by $\gamma_{A}, \gamma_{B}, \gamma_{C}$, and $\gamma_{D}$ the incircles of the quadrilaterals $A P O S, B Q O P$, $C R O Q$, and $D S O R$, respectively.

We start with proving that the quadrilateral $A B C D$ also has an incircle which will be referred to as $\Omega$. Denote the points of tangency as in Figure 1. It is well-known that $Q Q_{1}=O O_{1}$ (if $B C \| P R$, this is obvious; otherwise, one may regard the two circles involved as the incircle and an excircle of the triangle formed by the lines $O Q, P R$, and $B C$ ). Similarly, $O O_{1}=P P_{1}$. Hence we have $Q Q_{1}=P P_{1}$. The other equalities of segment lengths marked in Figure 1 can be proved analogously. These equalities, together with $A P_{1}=A S_{1}$ and similar ones, yield $A B+C D=A D+B C$, as required.


Figure 1

Next, let us draw the lines parallel to $Q S$ through $P$ and $R$, and also draw the lines parallel to $P R$ through $Q$ and $S$. These lines form a parallelogram; denote its vertices by $A^{\prime}, B^{\prime}, C^{\prime}$, and $D^{\prime}$ as shown in Figure 2.

Since the quadrilateral $A P O S$ has an incircle, we have $A P-A S=O P-O S=A^{\prime} S-A^{\prime} P$. It is well-known that in this case there also exists a circle $\omega_{A}$ tangent to the four rays $A P$, $A S, A^{\prime} P$, and $A^{\prime} S$. It is worth mentioning here that in case when, say, the lines $A B$ and $A^{\prime} B^{\prime}$ coincide, the circle $\omega_{A}$ is just tangent to $A B$ at $P$. We introduce the circles $\omega_{B}, \omega_{C}$, and $\omega_{D}$ in a similar manner.

Assume that the radii of the circles $\omega_{A}$ and $\omega_{C}$ are different. Let $X$ be the center of the homothety having a positive scale factor and mapping $\omega_{A}$ to $\omega_{C}$.

Now, Monge's theorem applied to the circles $\omega_{A}, \Omega$, and $\omega_{C}$ shows that the points $A, C$, and $X$ are collinear. Applying the same theorem to the circles $\omega_{A}, \omega_{B}$, and $\omega_{C}$, we see that the points $P, Q$, and $X$ are also collinear. Similarly, the points $R, S$, and $X$ are collinear, as required.

If the radii of $\omega_{A}$ and $\omega_{C}$ are equal but these circles do not coincide, then the degenerate version of the same theorem yields that the three lines $A C, P Q$, and $R S$ are parallel to the line of centers of $\omega_{A}$ and $\omega_{C}$.

Finally, we need to say a few words about the case when $\omega_{A}$ and $\omega_{C}$ coincide (and thus they also coincide with $\Omega, \omega_{B}$, and $\omega_{D}$ ). It may be regarded as the limit case in the following manner.


Figure 2

Let us fix the positions of $A, P, O$, and $S$ (thus we also fix the circles $\omega_{A}, \gamma_{A}, \gamma_{B}$, and $\gamma_{D}$ ). Now we vary the circle $\gamma_{C}$ inscribed into $\angle Q O R$; for each of its positions, one may reconstruct the lines $B C$ and $C D$ as the external common tangents to $\gamma_{B}, \gamma_{C}$ and $\gamma_{C}, \gamma_{D}$ different from $P R$ and $Q S$, respectively. After such variation, the circle $\Omega$ changes, so the result obtained above may be applied.

Solution 2. Applying Menelaus' theorem to $\triangle A B C$ with the line $P Q$ and to $\triangle A C D$ with the line $R S$, we see that the line $A C$ meets $P Q$ and $R S$ at the same point (possibly at infinity) if and only if

$$
\begin{equation*}
\frac{A P}{P B} \cdot \frac{B Q}{Q C} \cdot \frac{C R}{R D} \cdot \frac{D S}{S A}=1 \tag{1}
\end{equation*}
$$

So, it suffices to prove (1).
We start with the following result.
Lemma 1. Let $E F G H$ be a circumscribed quadrilateral, and let $M$ be its incenter. Then

$$
\frac{E F \cdot F G}{G H \cdot H E}=\frac{F M^{2}}{H M^{2}}
$$

Proof. Notice that $\angle E M H+\angle G M F=\angle F M E+\angle H M G=180^{\circ}, \angle F G M=\angle M G H$, and $\angle H E M=\angle M E F$ (see Figure 3). By the law of sines, we get

$$
\frac{E F}{F M} \cdot \frac{F G}{F M}=\frac{\sin \angle F M E \cdot \sin \angle G M F}{\sin \angle M E F \cdot \sin \angle F G M}=\frac{\sin \angle H M G \cdot \sin \angle E M H}{\sin \angle M G H \cdot \sin \angle H E M}=\frac{G H}{H M} \cdot \frac{H E}{H M} .
$$



Figure 3


Figure 4

We denote by $I, J, K$, and $L$ the incenters of the quadrilaterals $A P O S, B Q O P, C R O Q$, and $D S O R$, respectively. Applying Lemma 1 to these four quadrilaterals we get

$$
\frac{A P \cdot P O}{O S \cdot S A} \cdot \frac{B Q \cdot Q O}{O P \cdot P B} \cdot \frac{C R \cdot R O}{O Q \cdot Q C} \cdot \frac{D S \cdot S O}{O R \cdot R D}=\frac{P I^{2}}{S I^{2}} \cdot \frac{Q J^{2}}{P J^{2}} \cdot \frac{R K^{2}}{Q K^{2}} \cdot \frac{S L^{2}}{R L^{2}}
$$

which reduces to

$$
\begin{equation*}
\frac{A P}{P B} \cdot \frac{B Q}{Q C} \cdot \frac{C R}{R D} \cdot \frac{D S}{S A}=\frac{P I^{2}}{P J^{2}} \cdot \frac{Q J^{2}}{Q K^{2}} \cdot \frac{R K^{2}}{R L^{2}} \cdot \frac{S L^{2}}{S I^{2}} \tag{2}
\end{equation*}
$$

Next, we have $\angle I P J=\angle J O I=90^{\circ}$, and the line $O P$ separates $I$ and $J$ (see Figure 4). This means that the quadrilateral $I P J O$ is cyclic. Similarly, we get that the quadrilateral $J Q K O$ is cyclic with $\angle J Q K=90^{\circ}$. Thus, $\angle Q K J=\angle Q O J=\angle J O P=\angle J I P$. Hence, the right triangles $I P J$ and $K Q J$ are similar. Therefore, $\frac{P I}{P J}=\frac{Q K}{Q J}$. Likewise, we obtain $\frac{R K}{R L}=\frac{S I}{S L}$. These two equations together with (2) yield (1).
Comment. Instead of using the sine law, one may prove Lemma 1 by the following approach.


Figure 5
Let $N$ be the point such that $\triangle N H G \sim \triangle M E F$ and such that $N$ and $M$ lie on different sides of the line $G H$, as shown in Figure 5. Then $\angle G N H+\angle H M G=\angle F M E+\angle H M G=180^{\circ}$. So,
the quadrilateral $G N H M$ is cyclic. Thus, $\angle M N H=\angle M G H=\angle F G M$ and $\angle H M N=\angle H G N=$ $\angle E F M=\angle M F G$. Hence, $\triangle H M N \sim \triangle M F G$. Therefore, $\frac{H M}{H G}=\frac{H M}{H N} \cdot \frac{H N}{H G}=\frac{M F}{M G} \cdot \frac{E M}{E F}$. Similarly, we obtain $\frac{H M}{H E}=\frac{M F}{M E} \cdot \frac{G M}{G F}$. By multiplying these two equations, we complete the proof.

Solution 3. We present another approach for showing (1) from Solution 2.
Lemma 2. Let $E F G H$ and $E^{\prime} F^{\prime} G^{\prime} H^{\prime}$ be circumscribed quadrilaterals such that $\angle E+\angle E^{\prime}=$ $\angle F+\angle F^{\prime}=\angle G+\angle G^{\prime}=\angle H+\angle H^{\prime}=180^{\circ}$. Then

$$
\frac{E F \cdot G H}{F G \cdot H E}=\frac{E^{\prime} F^{\prime} \cdot G^{\prime} H^{\prime}}{F^{\prime} G^{\prime} \cdot H^{\prime} E^{\prime}}
$$

Proof. Let $M$ and $M^{\prime}$ be the incenters of $E F G H$ and $E^{\prime} F^{\prime} G^{\prime} H^{\prime}$, respectively. We use the notation [ $X Y Z$ ] for the area of a triangle $X Y Z$.

Taking into account the relation $\angle F M E+\angle F^{\prime} M^{\prime} E^{\prime}=180^{\circ}$ together with the analogous ones, we get

$$
\begin{aligned}
\frac{E F \cdot G H}{F G \cdot H E} & =\frac{[M E F] \cdot[M G H]}{[M F G] \cdot[M H E]}=\frac{M E \cdot M F \cdot \sin \angle F M E \cdot M G \cdot M H \cdot \sin \angle H M G}{M F \cdot M G \cdot \sin \angle G M F \cdot M H \cdot M E \cdot \sin \angle E M H} \\
& =\frac{M^{\prime} E^{\prime} \cdot M^{\prime} F^{\prime} \cdot \sin \angle F^{\prime} M^{\prime} E^{\prime} \cdot M^{\prime} G^{\prime} \cdot M^{\prime} H^{\prime} \cdot \sin \angle H^{\prime} M^{\prime} G^{\prime}}{M^{\prime} F^{\prime} \cdot M^{\prime} G^{\prime} \cdot \sin \angle G^{\prime} M^{\prime} F^{\prime} \cdot M^{\prime} H^{\prime} \cdot M^{\prime} E^{\prime} \cdot \sin \angle E^{\prime} M^{\prime} H^{\prime}}=\frac{E^{\prime} F^{\prime} \cdot G^{\prime} H^{\prime}}{F^{\prime} G^{\prime} \cdot H^{\prime} E^{\prime}} .
\end{aligned}
$$



Figure 6
Denote by $h$ the homothety centered at $O$ that maps the incircle of $C R O Q$ to the incircle of APOS. Let $Q^{\prime}=h(Q), C^{\prime}=h(C), R^{\prime}=h(R), O^{\prime}=O, S^{\prime}=S, A^{\prime}=A$, and $P^{\prime}=P$. Furthermore, define $B^{\prime}=A^{\prime} P^{\prime} \cap C^{\prime} Q^{\prime}$ and $D^{\prime}=A^{\prime} S^{\prime} \cap C^{\prime} R^{\prime}$ as shown in Figure 6. Then

$$
\frac{A P \cdot O S}{P O \cdot S A}=\frac{A^{\prime} P^{\prime} \cdot O^{\prime} S^{\prime}}{P^{\prime} O^{\prime} \cdot S^{\prime} A^{\prime}}
$$

holds trivially. We also have

$$
\frac{C R \cdot O Q}{R O \cdot Q C}=\frac{C^{\prime} R^{\prime} \cdot O^{\prime} Q^{\prime}}{R^{\prime} O^{\prime} \cdot Q^{\prime} C^{\prime}}
$$

by the similarity of the quadrilaterals $C R O Q$ and $C^{\prime} R^{\prime} O^{\prime} Q^{\prime}$.

Next, consider the circumscribed quadrilaterals $B Q O P$ and $B^{\prime} Q^{\prime} O^{\prime} P^{\prime}$ whose incenters lie on different sides of the quadrilaterals' shared side line $O P=O^{\prime} P^{\prime}$. Observe that $B Q \| B^{\prime} Q^{\prime}$ and that $B^{\prime}$ and $Q^{\prime}$ lie on the lines $B P$ and $Q O$, respectively. It is now easy to see that the two quadrilaterals satisfy the hypotheses of Lemma 2. Thus, we deduce

$$
\frac{B Q \cdot O P}{Q O \cdot P B}=\frac{B^{\prime} Q^{\prime} \cdot O^{\prime} P^{\prime}}{Q^{\prime} O^{\prime} \cdot P^{\prime} B^{\prime}}
$$

Similarly, we get

$$
\frac{D S \cdot O R}{S O \cdot R D}=\frac{D^{\prime} S^{\prime} \cdot O^{\prime} R^{\prime}}{S^{\prime} O^{\prime} \cdot R^{\prime} D^{\prime}} .
$$

Multiplying these four equations, we obtain

$$
\begin{equation*}
\frac{A P}{P B} \cdot \frac{B Q}{Q C} \cdot \frac{C R}{R D} \cdot \frac{D S}{S A}=\frac{A^{\prime} P^{\prime}}{P^{\prime} B^{\prime}} \cdot \frac{B^{\prime} Q^{\prime}}{Q^{\prime} C^{\prime}} \cdot \frac{C^{\prime} R^{\prime}}{R^{\prime} D^{\prime}} \cdot \frac{D^{\prime} S^{\prime}}{S^{\prime} A^{\prime}} \tag{3}
\end{equation*}
$$

Finally, we apply Brianchon's theorem to the circumscribed hexagon $A^{\prime} P^{\prime} R^{\prime} C^{\prime} Q^{\prime} S^{\prime}$ and deduce that the lines $A^{\prime} C^{\prime}, P^{\prime} Q^{\prime}$, and $R^{\prime} S^{\prime}$ are either concurrent or parallel to each other. So, by Menelaus' theorem, we obtain

$$
\frac{A^{\prime} P^{\prime}}{P^{\prime} B^{\prime}} \cdot \frac{B^{\prime} Q^{\prime}}{Q^{\prime} C^{\prime}} \cdot \frac{C^{\prime} R^{\prime}}{R^{\prime} D^{\prime}} \cdot \frac{D^{\prime} S^{\prime}}{S^{\prime} A^{\prime}}=1
$$

This equation together with (3) yield (1).

G8. A triangulation of a convex polygon $\Pi$ is a partitioning of $\Pi$ into triangles by diagonals having no common points other than the vertices of the polygon. We say that a triangulation is a Thaiangulation if all triangles in it have the same area.

Prove that any two different Thaiangulations of a convex polygon $\Pi$ differ by exactly two triangles. (In other words, prove that it is possible to replace one pair of triangles in the first Thaiangulation with a different pair of triangles so as to obtain the second Thaiangulation.)
(Bulgaria)
Solution 1. We denote by $[S]$ the area of a polygon $S$.
Recall that each triangulation of a convex $n$-gon has exactly $n-2$ triangles. This means that all triangles in any two Thaiangulations of a convex polygon $\Pi$ have the same area.

Let $\mathcal{T}$ be a triangulation of a convex polygon $\Pi$. If four vertices $A, B, C$, and $D$ of $\Pi$ form a parallelogram, and $\mathcal{T}$ contains two triangles whose union is this parallelogram, then we say that $\mathcal{T}$ contains parallelogram $A B C D$. Notice here that if two Thaiangulations $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ of $\Pi$ differ by two triangles, then the union of these triangles is a quadrilateral each of whose diagonals bisects its area, i.e., a parallelogram.

We start with proving two properties of triangulations.
Lemma 1. A triangulation of a convex polygon $\Pi$ cannot contain two parallelograms.
Proof. Arguing indirectly, assume that $P_{1}$ and $P_{2}$ are two parallelograms contained in some triangulation $\mathcal{T}$. If they have a common triangle in $\mathcal{T}$, then we may assume that $P_{1}$ consists of triangles $A B C$ and $A D C$ of $\mathcal{T}$, while $P_{2}$ consists of triangles $A D C$ and $C D E$ (see Figure 1). But then $B C\|A D\| C E$, so the three vertices $B, C$, and $E$ of $\Pi$ are collinear, which is absurd.

Assume now that $P_{1}$ and $P_{2}$ contain no common triangle. Let $P_{1}=A B C D$. The sides $A B$, $B C, C D$, and $D A$ partition $\Pi$ into several parts, and $P_{2}$ is contained in one of them; we may assume that this part is cut off from $P_{1}$ by $A D$. Then one may label the vertices of $P_{2}$ by $X$, $Y, Z$, and $T$ so that the polygon $A B C D X Y Z T$ is convex (see Figure 2; it may happen that $D=X$ and/or $T=A$, but still this polygon has at least six vertices). But the sum of the external angles of this polygon at $B, C, Y$, and $Z$ is already $360^{\circ}$, which is impossible. A final contradiction.


Figure 1


Figure 2


Figure 3

Lemma 2. Every triangle in a Thaiangulation $\mathcal{T}$ of $\Pi$ contains a side of $\Pi$.
Proof. Let $A B C$ be a triangle in $\mathcal{T}$. Apply an affine transform such that $A B C$ maps to an equilateral triangle; let $A^{\prime} B^{\prime} C^{\prime}$ be the image of this triangle, and $\Pi^{\prime}$ be the image of $\Pi$. Clearly, $\mathcal{T}$ maps into a Thaiangulation $\mathcal{T}^{\prime}$ of $\Pi^{\prime}$.

Assume that none of the sides of $\triangle A^{\prime} B^{\prime} C^{\prime}$ is a side of $\Pi^{\prime}$. Then $\mathcal{T}^{\prime}$ contains some other triangles with these sides, say, $A^{\prime} B^{\prime} Z, C^{\prime} A^{\prime} Y$, and $B^{\prime} C^{\prime} X$; notice that $A^{\prime} Z B^{\prime} X C^{\prime} Y$ is a convex hexagon (see Figure 3). The sum of its external angles at $X, Y$, and $Z$ is less than $360^{\circ}$. So one of these angles (say, at $Z$ ) is less than $120^{\circ}$, hence $\angle A^{\prime} Z B^{\prime}>60^{\circ}$. Then $Z$ lies on a circular arc subtended by $A^{\prime} B^{\prime}$ and having angular measure less than $240^{\circ}$; consequently, the altitude $Z H$ of $\triangle A^{\prime} B^{\prime} Z$ is less than $\sqrt{3} A^{\prime} B^{\prime} / 2$. Thus $\left[A^{\prime} B^{\prime} Z\right]<\left[A^{\prime} B^{\prime} C^{\prime}\right]$, and $\mathcal{T}^{\prime}$ is not a Thaiangulation. A contradiction.

Now we pass to the solution. We say that a triangle in a triangulation of $\Pi$ is an ear if it contains two sides of $\Pi$. Note that each triangulation of a polygon contains some ear.

Arguing indirectly, we choose a convex polygon $\Pi$ with the least possible number of sides such that some two Thaiangulations $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ of $\Pi$ violate the problem statement (thus $\Pi$ has at least five sides). Consider now any ear $A B C$ in $\mathcal{T}_{1}$, with $A C$ being a diagonal of $\Pi$. If $\mathcal{T}_{2}$ also contains $\triangle A B C$, then one may cut $\triangle A B C$ off from $\Pi$, getting a polygon with a smaller number of sides which also violates the problem statement. This is impossible; thus $\mathcal{T}_{2}$ does not contain $\triangle A B C$.

Next, $\mathcal{T}_{1}$ contains also another triangle with side $A C$, say $\triangle A C D$. By Lemma 2, this triangle contains a side of $\Pi$, so $D$ is adjacent to either $A$ or $C$ on the boundary of $\Pi$. We may assume that $D$ is adjacent to $C$.

Assume that $\mathcal{T}_{2}$ does not contain the triangle $B C D$. Then it contains two different triangles $B C X$ and $C D Y$ (possibly, with $X=Y$ ); since these triangles have no common interior points, the polygon $A B C D Y X$ is convex (see Figure 4). But, since $[A B C]=[B C X]=$ $[A C D]=[C D Y]$, we get $A X \| B C$ and $A Y \| C D$ which is impossible. Thus $\mathcal{T}_{2}$ contains $\triangle B C D$.

Therefore, $[A B D]=[A B C]+[A C D]-[B C D]=[A B C]$, and $A B C D$ is a parallelogram contained in $\mathcal{T}_{1}$. Let $\mathcal{T}^{\prime}$ be the Thaiangulation of $\Pi$ obtained from $\mathcal{T}_{1}$ by replacing the diagonal $A C$ with $B D$; then $\mathcal{T}^{\prime}$ is distinct from $\mathcal{T}_{2}$ (otherwise $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ would differ by two triangles). Moreover, $\mathcal{T}^{\prime}$ shares a common ear $B C D$ with $\mathcal{T}_{2}$. As above, cutting this ear away we obtain that $\mathcal{T}_{2}$ and $\mathcal{T}^{\prime}$ differ by two triangles forming a parallelogram different from $A B C D$. Thus $\mathcal{T}^{\prime}$ contains two parallelograms, which contradicts Lemma 1.


Figure 4


Figure 5

Comment 1. Lemma 2 is equivalent to the well-known Erdős-Debrunner inequality stating that for any triangle $P Q R$ and any points $A, B, C$ lying on the sides $Q R, R P$, and $P Q$, respectively, we have

$$
\begin{equation*}
[A B C] \geqslant \min \{[A B R],[B C P],[C A Q]\} \tag{1}
\end{equation*}
$$

To derive this inequality from Lemma 2, one may assume that (1) does not hold, and choose some points $X, Y$, and $Z$ inside the triangles $B C P, C A Q$, and $A B R$, respectively, so that $[A B C]=$ $[A B Z]=[B C X]=[C A Y]$. Then a convex hexagon $A Z B X C Y$ has a Thaiangulation containing $\triangle A B C$, which contradicts Lemma 2.

Conversely, assume that a Thaiangulation $\mathcal{T}$ of $\Pi$ contains a triangle $A B C$ none of whose sides is a side of $\Pi$, and let $A B Z, A Y C$, and $X B C$ be other triangles in $\mathcal{T}$ containing the corresponding sides. Then $A Z B X C Y$ is a convex hexagon.

Consider the lines through $A, B$, and $C$ parallel to $Y Z, Z X$, and $X Y$, respectively. They form a triangle $X^{\prime} Y^{\prime} Z^{\prime}$ similar to $\triangle X Y Z$ (see Figure 5). By (1) we have

$$
[A B C] \geqslant \min \left\{\left[A B Z^{\prime}\right],\left[B C X^{\prime}\right],\left[C A Y^{\prime}\right]\right\}>\min \{[A B Z],[B C X],[C A Y]\}
$$

Solution 2. We will make use of the preliminary observations from Solution 1, together with Lemma 1.

Arguing indirectly, we choose a convex polygon $\Pi$ with the least possible number of sides such that some two Thaiangulations $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ of $\Pi$ violate the statement (thus $\Pi$ has at least five sides). Assume that $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ share a diagonal $d$ splitting $\Pi$ into two smaller polygons $\Pi_{1}$ and $\Pi_{2}$. Since the problem statement holds for any of them, the induced Thaiangulations of each of $\Pi_{i}$ differ by two triangles forming a parallelogram (the Thaiangulations induced on $\Pi_{i}$ by $\mathcal{T}_{1}$ and $T_{2}$ may not coincide, otherwise $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ would differ by at most two triangles). But both these parallelograms are contained in $\mathcal{T}_{1}$; this contradicts Lemma 1. Therefore, $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ share no diagonal. Hence they also share no triangle.

We consider two cases.
Case 1. Assume that some vertex $B$ of $\Pi$ is an endpoint of some diagonal in $\mathcal{T}_{1}$, as well as an endpoint of some diagonal in $\mathcal{T}_{2}$.

Let $A$ and $C$ be the vertices of $\Pi$ adjacent to $B$. Then $\mathcal{T}_{1}$ contains some triangles $A B X$ and $B C Y$, while $\mathcal{T}_{2}$ contains some triangles $A B X^{\prime}$ and $B C Y^{\prime}$. Here, some of the points $X$, $X^{\prime}, Y$, and $Y^{\prime}$ may coincide; however, in view of our assumption together with the fact that $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$ share no triangle, all four triangles $A B X, B C Y, A B X^{\prime}$, and $B C Y^{\prime}$ are distinct.

Since $[A B X]=[B C Y]=\left[A B X^{\prime}\right]=\left[B C Y^{\prime}\right]$, we have $X X^{\prime} \| A B$ and $Y Y^{\prime} \| B C$. Now, if $X=Y$, then $X^{\prime}$ and $Y^{\prime}$ lie on different lines passing through $X$ and are distinct from that point, so that $X^{\prime} \neq Y^{\prime}$. In this case, we may switch the two Thaiangulations. So, hereafter we assume that $X \neq Y$.

In the convex pentagon $A B C Y X$ we have either $\angle B A X+\angle A X Y>180^{\circ}$ or $\angle X Y C+$ $\angle Y C B>180^{\circ}$ (or both); due to the symmetry, we may assume that the first inequality holds. Let $r$ be the ray emerging from $X$ and co-directed with $\overrightarrow{A B}$; our inequality shows that $r$ points to the interior of the pentagon (and thus to the interior of $\Pi$ ). Therefore, the ray opposite to $r$ points outside $\Pi$, so $X^{\prime}$ lies on $r$; moreover, $X^{\prime}$ lies on the "arc" $C Y$ of $\Pi$ not containing $X$. So the segments $X X^{\prime}$ and $Y B$ intersect (see Figure 6).

Let $O$ be the intersection point of the rays $r$ and $B C$. Since the triangles $A B X^{\prime}$ and $B C Y^{\prime}$ have no common interior points, $Y^{\prime}$ must lie on the "arc" $C X^{\prime}$ which is situated inside the triangle $X B O$. Therefore, the line $Y Y^{\prime}$ meets two sides of $\triangle X B O$, none of which may be $X B$ (otherwise the diagonals $X B$ and $Y Y^{\prime}$ would share a common point). Thus $Y Y^{\prime}$ intersects $B O$, which contradicts $Y Y^{\prime} \| B C$.


Figure 6

Case 2. In the remaining case, each vertex of $\Pi$ is an endpoint of a diagonal in at most one of $\mathcal{T}_{1}$ and $\mathcal{T}_{2}$. On the other hand, a triangulation cannot contain two consecutive vertices with no diagonals from each. Therefore, the vertices of $\Pi$ alternatingly emerge diagonals in $\mathcal{T}_{1}$ and in $\mathcal{T}_{2}$. In particular, $\Pi$ has an even number of sides.

Next, we may choose five consecutive vertices $A, B, C, D$, and $E$ of $\Pi$ in such a way that

$$
\begin{equation*}
\angle A B C+\angle B C D>180^{\circ} \text { and } \angle B C D+\angle C D E>180^{\circ} . \tag{2}
\end{equation*}
$$

In order to do this, it suffices to choose three consecutive vertices $B, C$, and $D$ of $\Pi$ such that the sum of their external angles is at most $180^{\circ}$. This is possible, since $\Pi$ has at least six sides.


Figure 7
We may assume that $\mathcal{T}_{1}$ has no diagonals from $B$ and $D$ (and thus contains the triangles $A B C$ and $C D E$ ), while $\mathcal{T}_{2}$ has no diagonals from $A, C$, and $E$ (and thus contains the triangle $B C D)$. Now, since $[A B C]=[B C D]=[C D E]$, we have $A D \| B C$ and $B E \| C D$ (see Figure 7). By (2) this yields that $A D>B C$ and $B E>C D$. Let $X=A C \cap B D$ and $Y=C E \cap B D$; then the inequalities above imply that $A X>C X$ and $E Y>C Y$.

Finally, $\mathcal{T}_{2}$ must also contain some triangle $B D Z$ with $Z \neq C$; then the ray $C Z$ lies in the angle $A C E$. Since $[B C D]=[B D Z]$, the diagonal $B D$ bisects $C Z$. Together with the inequalities above, this yields that $Z$ lies inside the triangle $A C E$ (but $Z$ is distinct from $A$ and $E$ ), which is impossible. The final contradiction.

Comment 2. Case 2 may also be accomplished with the use of Lemma 2. Indeed, since each triangulation of an $n$-gon contains $n-2$ triangles neither of which may contain three sides of $\Pi$, Lemma 2 yields that each Thaiangulation contains exactly two ears. But each vertex of $\Pi$ is a vertex of an ear either in $\mathcal{T}_{1}$ or in $\mathcal{T}_{2}$, so $\Pi$ cannot have more than four vertices.

## Number Theory

N1. Determine all positive integers $M$ for which the sequence $a_{0}, a_{1}, a_{2}, \ldots$, defined by $a_{0}=\frac{2 M+1}{2}$ and $a_{k+1}=a_{k}\left\lfloor a_{k}\right\rfloor$ for $k=0,1,2, \ldots$, contains at least one integer term.
(Luxembourg)
Answer. All integers $M \geqslant 2$.
Solution 1. Define $b_{k}=2 a_{k}$ for all $k \geqslant 0$. Then

$$
b_{k+1}=2 a_{k+1}=2 a_{k}\left\lfloor a_{k}\right\rfloor=b_{k}\left\lfloor\frac{b_{k}}{2}\right\rfloor .
$$

Since $b_{0}$ is an integer, it follows that $b_{k}$ is an integer for all $k \geqslant 0$.
Suppose that the sequence $a_{0}, a_{1}, a_{2}, \ldots$ does not contain any integer term. Then $b_{k}$ must be an odd integer for all $k \geqslant 0$, so that

$$
\begin{equation*}
b_{k+1}=b_{k}\left\lfloor\frac{b_{k}}{2}\right\rfloor=\frac{b_{k}\left(b_{k}-1\right)}{2} . \tag{1}
\end{equation*}
$$

Hence

$$
\begin{equation*}
b_{k+1}-3=\frac{b_{k}\left(b_{k}-1\right)}{2}-3=\frac{\left(b_{k}-3\right)\left(b_{k}+2\right)}{2} \tag{2}
\end{equation*}
$$

for all $k \geqslant 0$.
Suppose that $b_{0}-3>0$. Then equation (2) yields $b_{k}-3>0$ for all $k \geqslant 0$. For each $k \geqslant 0$, define $c_{k}$ to be the highest power of 2 that divides $b_{k}-3$. Since $b_{k}-3$ is even for all $k \geqslant 0$, the number $c_{k}$ is positive for every $k \geqslant 0$.

Note that $b_{k}+2$ is an odd integer. Therefore, from equation (2), we have that $c_{k+1}=c_{k}-1$. Thus, the sequence $c_{0}, c_{1}, c_{2}, \ldots$ of positive integers is strictly decreasing, a contradiction. So, $b_{0}-3 \leqslant 0$, which implies $M=1$.

For $M=1$, we can check that the sequence is constant with $a_{k}=\frac{3}{2}$ for all $k \geqslant 0$. Therefore, the answer is $M \geqslant 2$.

Solution 2. We provide an alternative way to show $M=1$ once equation (1) has been reached. We claim that $b_{k} \equiv 3\left(\bmod 2^{m}\right)$ for all $k \geqslant 0$ and $m \geqslant 1$. If this is true, then we would have $b_{k}=3$ for all $k \geqslant 0$ and hence $M=1$.

To establish our claim, we proceed by induction on $m$. The base case $b_{k} \equiv 3(\bmod 2)$ is true for all $k \geqslant 0$ since $b_{k}$ is odd. Now suppose that $b_{k} \equiv 3\left(\bmod 2^{m}\right)$ for all $k \geqslant 0$. Hence $b_{k}=2^{m} d_{k}+3$ for some integer $d_{k}$. We have

$$
3 \equiv b_{k+1} \equiv\left(2^{m} d_{k}+3\right)\left(2^{m-1} d_{k}+1\right) \equiv 3 \cdot 2^{m-1} d_{k}+3 \quad\left(\bmod 2^{m}\right)
$$

so that $d_{k}$ must be even. This implies that $b_{k} \equiv 3\left(\bmod 2^{m+1}\right)$, as required.
Comment. The reason the number 3 which appears in both solutions is important, is that it is a nontrivial fixed point of the recurrence relation for $b_{k}$.

N2. Let $a$ and $b$ be positive integers such that $a!b!$ is a multiple of $a!+b!$. Prove that $3 a \geqslant 2 b+2$.
(United Kingdom)
Solution 1. If $a>b$, we immediately get $3 a \geqslant 2 b+2$. In the case $a=b$, the required inequality is equivalent to $a \geqslant 2$, which can be checked easily since $(a, b)=(1,1)$ does not satisfy $a!+b!\mid a!b!$. We now assume $a<b$ and denote $c=b-a$. The required inequality becomes $a \geqslant 2 c+2$.

Suppose, to the contrary, that $a \leqslant 2 c+1$. Define $M=\frac{b!}{a!}=(a+1)(a+2) \cdots(a+c)$. Since $a!+b!\mid a!b!$ implies $1+M \mid a!M$, we obtain $1+M \mid a!$. Note that we must have $c<a$; otherwise $1+M>a!$, which is impossible. We observe that $c!\mid M$ since $M$ is a product of $c$ consecutive integers. Thus $\operatorname{gcd}(1+M, c!)=1$, which implies

$$
\begin{equation*}
1+M \left\lvert\, \frac{a!}{c!}=(c+1)(c+2) \cdots a\right. \tag{1}
\end{equation*}
$$

If $a \leqslant 2 c$, then $\frac{a!}{c!}$ is a product of $a-c \leqslant c$ integers not exceeding $a$ whereas $M$ is a product of $c$ integers exceeding $a$. Therefore, $1+M>\frac{a!}{c!}$, which is a contradiction.

It remains to exclude the case $a=2 c+1$. Since $a+1=2(c+1)$, we have $c+1 \mid M$. Hence, we can deduce from (1) that $1+M \mid(c+2)(c+3) \cdots a$. Now $(c+2)(c+3) \cdots a$ is a product of $a-c-1=c$ integers not exceeding $a$; thus it is smaller than $1+M$. Again, we arrive at a contradiction.

Comment 1. One may derive a weaker version of (1) and finish the problem as follows. After assuming $a \leqslant 2 c+1$, we have $\left\lfloor\frac{a}{2}\right\rfloor \leqslant c$, so $\left.\left\lfloor\frac{a}{2}\right\rfloor!\right\rvert\, M$. Therefore,

$$
1+M \left\lvert\,\left(\left\lfloor\frac{a}{2}\right\rfloor+1\right)\left(\left\lfloor\frac{a}{2}\right\rfloor+2\right) \cdots a\right.
$$

Observe that $\left(\left\lfloor\frac{a}{2}\right\rfloor+1\right)\left(\left\lfloor\frac{a}{2}\right\rfloor+2\right) \cdots a$ is a product of $\left\lceil\frac{a}{2}\right\rceil$ integers not exceeding $a$. This leads to a contradiction when $a$ is even since $\left\lceil\frac{a}{2}\right\rceil=\frac{a}{2} \leqslant c$ and $M$ is a product of $c$ integers exceeding $a$.

When $a$ is odd, we can further deduce that $1+M \left\lvert\,\left(\frac{a+3}{2}\right)\left(\frac{a+5}{2}\right) \cdots a\right.$ since $\left.\left\lfloor\frac{a}{2}\right\rfloor+1=\frac{a+1}{2} \right\rvert\, a+1$. Now $\left(\frac{a+3}{2}\right)\left(\frac{a+5}{2}\right) \cdots a$ is a product of $\frac{a-1}{2} \leqslant c$ numbers not exceeding $a$, and we get a contradiction.

Solution 2. As in Solution 1, we may assume that $a<b$ and let $c=b-a$. Suppose, to the contrary, that $a \leqslant 2 c+1$. From $a!+b!\mid a!b!$, we have

$$
N=1+(a+1)(a+2) \cdots(a+c) \mid(a+c)!,
$$

which implies that all prime factors of $N$ are at most $a+c$.
Let $p$ be a prime factor of $N$. If $p \leqslant c$ or $p \geqslant a+1$, then $p$ divides one of $a+1, \ldots, a+c$ which is impossible. Hence $a \geqslant p \geqslant c+1$. Furthermore, we must have $2 p>a+c$; otherwise, $a+1 \leqslant 2 c+2 \leqslant 2 p \leqslant a+c$ so $p \mid N-1$, again impossible. Thus, we have $p \in\left(\frac{a+c}{2}, a\right]$, and $p^{2} \nmid(a+c)$ ! since $2 p>a+c$. Therefore, $p^{2} \nmid N$ as well.

If $a \leqslant c+2$, then the interval $\left(\frac{a+c}{2}, a\right]$ contains at most one integer and hence at most one prime number, which has to be $a$. Since $p^{2} \nmid N$, we must have $N=p=a$ or $N=1$, which is absurd since $N>a \geqslant 1$. Thus, we have $a \geqslant c+3$, and so $\frac{a+c+1}{2} \geqslant c+2$. It follows that $p$ lies in the interval $[c+2, a]$.

Thus, every prime appearing in the prime factorization of $N$ lies in the interval $[c+2, a]$, and its exponent is exactly 1 . So we must have $N \mid(c+2)(c+3) \cdots a$. However, $(c+2)(c+3) \cdots a$ is a product of $a-c-1 \leqslant c$ numbers not exceeding $a$, so it is less than $N$. This is a contradiction.

Comment 2. The original problem statement also asks to determine when the equality $3 a=2 b+2$ holds. It can be checked that the answer is $(a, b)=(2,2),(4,5)$.

N3. Let $m$ and $n$ be positive integers such that $m>n$. Define $x_{k}=(m+k) /(n+k)$ for $k=$ $1,2, \ldots, n+1$. Prove that if all the numbers $x_{1}, x_{2}, \ldots, x_{n+1}$ are integers, then $x_{1} x_{2} \cdots x_{n+1}-1$ is divisible by an odd prime.
(Austria)
Solution. Assume that $x_{1}, x_{2}, \ldots, x_{n+1}$ are integers. Define the integers

$$
a_{k}=x_{k}-1=\frac{m+k}{n+k}-1=\frac{m-n}{n+k}>0
$$

for $k=1,2, \ldots, n+1$.
Let $P=x_{1} x_{2} \cdots x_{n+1}-1$. We need to prove that $P$ is divisible by an odd prime, or in other words, that $P$ is not a power of 2 . To this end, we investigate the powers of 2 dividing the numbers $a_{k}$.

Let $2^{d}$ be the largest power of 2 dividing $m-n$, and let $2^{c}$ be the largest power of 2 not exceeding $2 n+1$. Then $2 n+1 \leqslant 2^{c+1}-1$, and so $n+1 \leqslant 2^{c}$. We conclude that $2^{c}$ is one of the numbers $n+1, n+2, \ldots, 2 n+1$, and that it is the only multiple of $2^{c}$ appearing among these numbers. Let $\ell$ be such that $n+\ell=2^{c}$. Since $\frac{m-n}{n+\ell}$ is an integer, we have $d \geqslant c$. Therefore, $2^{d-c+1} \nmid a_{\ell}=\frac{m-n}{n+\ell}$, while $2^{d-c+1} \mid a_{k}$ for all $k \in\{1, \ldots, n+1\} \backslash\{\ell\}$.

Computing modulo $2^{d-c+1}$, we get

$$
P=\left(a_{1}+1\right)\left(a_{2}+1\right) \cdots\left(a_{n+1}+1\right)-1 \equiv\left(a_{\ell}+1\right) \cdot 1^{n}-1 \equiv a_{\ell} \not \equiv 0 \quad\left(\bmod 2^{d-c+1}\right) .
$$

Therefore, $2^{d-c+1} \nmid P$.
On the other hand, for any $k \in\{1, \ldots, n+1\} \backslash\{\ell\}$, we have $2^{d-c+1} \mid a_{k}$. So $P \geqslant a_{k} \geqslant 2^{d-c+1}$, and it follows that $P$ is not a power of 2 .

Comment. Instead of attempting to show that $P$ is not a power of 2 , one may try to find an odd factor of $P$ (greater than 1 ) as follows:

From $a_{k}=\frac{m-n}{n+k} \in \mathbb{Z}_{>0}$, we get that $m-n$ is divisible by $n+1, n+2, \ldots, 2 n+1$, and thus it is also divisible by their least common multiple $L$. So $m-n=q L$ for some positive integer $q$; hence $x_{k}=q \cdot \frac{L}{n+k}+1$.

Then, since $n+1 \leqslant 2^{c}=n+\ell \leqslant 2 n+1 \leqslant 2^{c+1}-1$, we have $2^{c} \mid L$, but $2^{c+1} \nmid L$. So $\frac{L}{n+\ell}$ is odd, while $\frac{L}{n+k}$ is even for $k \neq \ell$. Computing modulo $2 q$ yields

$$
x_{1} x_{2} \cdots x_{n+1}-1 \equiv(q+1) \cdot 1^{n}-1 \equiv q \quad(\bmod 2 q) .
$$

Thus, $x_{1} x_{2} \cdots x_{n+1}-1=2 q r+q=q(2 r+1)$ for some integer $r$.
Since $x_{1} x_{2} \cdots x_{n+1}-1 \geqslant x_{1} x_{2}-1 \geqslant(q+1)^{2}-1>q$, we have $r \geqslant 1$. This implies that $x_{1} x_{2} \cdots x_{n+1}-1$ is divisible by an odd prime.

N4. Suppose that $a_{0}, a_{1}, \ldots$ and $b_{0}, b_{1}, \ldots$ are two sequences of positive integers satisfying $a_{0}, b_{0} \geqslant 2$ and

$$
a_{n+1}=\operatorname{gcd}\left(a_{n}, b_{n}\right)+1, \quad b_{n+1}=\operatorname{lcm}\left(a_{n}, b_{n}\right)-1
$$

for all $n \geqslant 0$. Prove that the sequence $\left(a_{n}\right)$ is eventually periodic; in other words, there exist integers $N \geqslant 0$ and $t>0$ such that $a_{n+t}=a_{n}$ for all $n \geqslant N$.
(France)
Solution 1. Let $s_{n}=a_{n}+b_{n}$. Notice that if $a_{n} \mid b_{n}$, then $a_{n+1}=a_{n}+1, b_{n+1}=b_{n}-1$ and $s_{n+1}=s_{n}$. So, $a_{n}$ increases by 1 and $s_{n}$ does not change until the first index is reached with $a_{n} \nmid s_{n}$. Define

$$
W_{n}=\left\{m \in \mathbb{Z}_{>0}: m \geqslant a_{n} \text { and } m \nmid s_{n}\right\} \quad \text { and } \quad w_{n}=\min W_{n}
$$

Claim 1. The sequence $\left(w_{n}\right)$ is non-increasing.
Proof. If $a_{n} \mid b_{n}$ then $a_{n+1}=a_{n}+1$. Due to $a_{n} \mid s_{n}$, we have $a_{n} \notin W_{n}$. Moreover $s_{n+1}=s_{n}$; therefore, $W_{n+1}=W_{n}$ and $w_{n+1}=w_{n}$.

Otherwise, if $a_{n} \nmid b_{n}$, then $a_{n} \nmid s_{n}$, so $a_{n} \in W_{n}$ and thus $w_{n}=a_{n}$. We show that $a_{n} \in W_{n+1}$; this implies $w_{n+1} \leqslant a_{n}=w_{n}$. By the definition of $W_{n+1}$, we need that $a_{n} \geqslant a_{n+1}$ and $a_{n} \nmid s_{n+1}$. The first relation holds because of $\operatorname{gcd}\left(a_{n}, b_{n}\right)<a_{n}$. For the second relation, observe that in $s_{n+1}=\operatorname{gcd}\left(a_{n}, b_{n}\right)+\operatorname{lcm}\left(a_{n}, b_{n}\right)$, the second term is divisible by $a_{n}$, but the first term is not. So $a_{n} \nmid s_{n+1}$; that completes the proof of the claim.

Let $w=\min _{n} w_{n}$ and let $N$ be an index with $w=w_{N}$. Due to Claim 1, we have $w_{n}=w$ for all $n \geqslant N$.

Let $g_{n}=\operatorname{gcd}\left(w, s_{n}\right)$. As we have seen, starting from an arbitrary index $n \geqslant N$, the sequence $a_{n}, a_{n+1}, \ldots$ increases by 1 until it reaches $w$, which is the first value not dividing $s_{n}$; then it drops to $\operatorname{gcd}\left(w, s_{n}\right)+1=g_{n}+1$.
Claim 2. The sequence $\left(g_{n}\right)$ is constant for $n \geqslant N$.
Proof. If $a_{n} \mid b_{n}$, then $s_{n+1}=s_{n}$ and hence $g_{n+1}=g_{n}$. Otherwise we have $a_{n}=w$,

$$
\begin{align*}
\operatorname{gcd}\left(a_{n}, b_{n}\right) & =\operatorname{gcd}\left(a_{n}, s_{n}\right)=\operatorname{gcd}\left(w, s_{n}\right)=g_{n} \\
s_{n+1} & =\operatorname{gcd}\left(a_{n}, b_{n}\right)+\operatorname{lcm}\left(a_{n}, b_{n}\right)=g_{n}+\frac{a_{n} b_{n}}{g_{n}}=g_{n}+\frac{w\left(s_{n}-w\right)}{g_{n}}  \tag{1}\\
\text { and } \quad g_{n+1} & =\operatorname{gcd}\left(w, s_{n+1}\right)=\operatorname{gcd}\left(w, g_{n}+\frac{s_{n}-w}{g_{n}} w\right)=\operatorname{gcd}\left(w, g_{n}\right)=g_{n}
\end{align*}
$$

Let $g=g_{N}$. We have proved that the sequence $\left(a_{n}\right)$ eventually repeats the following cycle:

$$
g+1 \mapsto g+2 \mapsto \ldots \mapsto w \mapsto g+1
$$

Solution 2. By Claim 1 in the first solution, we have $a_{n} \leqslant w_{n} \leqslant w_{0}$, so the sequence $\left(a_{n}\right)$ is bounded, and hence it has only finitely many values.

Let $M=\operatorname{lcm}\left(a_{1}, a_{2}, \ldots\right)$, and consider the sequence $b_{n}$ modulo $M$. Let $r_{n}$ be the remainder of $b_{n}$, divided by $M$. For every index $n$, since $a_{n}|M| b_{n}-r_{n}$, we have $\operatorname{gcd}\left(a_{n}, b_{n}\right)=\operatorname{gcd}\left(a_{n}, r_{n}\right)$, and therefore

$$
a_{n+1}=\operatorname{gcd}\left(a_{n}, r_{n}\right)+1 .
$$

Moreover,

$$
\begin{aligned}
r_{n+1} & \equiv b_{n+1}=\operatorname{lcm}\left(a_{n}, b_{n}\right)-1=\frac{a_{n}}{\operatorname{gcd}\left(a_{n}, b_{n}\right)} b_{n}-1 \\
& =\frac{a_{n}}{\operatorname{gcd}\left(a_{n}, r_{n}\right)} b_{n}-1 \equiv \frac{a_{n}}{\operatorname{gcd}\left(a_{n}, r_{n}\right)} r_{n}-1 \quad(\bmod M) .
\end{aligned}
$$

Hence, the pair $\left(a_{n}, r_{n}\right)$ uniquely determines the pair $\left(a_{n+1}, r_{n+1}\right)$. Since there are finitely many possible pairs, the sequence of pairs $\left(a_{n}, r_{n}\right)$ is eventually periodic; in particular, the sequence $\left(a_{n}\right)$ is eventually periodic.

Comment. We show that there are only four possibilities for $g$ and $w$ (as defined in Solution 1), namely

$$
\begin{equation*}
(w, g) \in\{(2,1),(3,1),(4,2),(5,1)\} . \tag{2}
\end{equation*}
$$

This means that the sequence $\left(a_{n}\right)$ eventually repeats one of the following cycles:

$$
\begin{equation*}
(2), \quad(2,3), \quad(3,4), \quad \text { or } \quad(2,3,4,5) . \tag{3}
\end{equation*}
$$

Using the notation of Solution 1, for $n \geqslant N$ the sequence $\left(a_{n}\right)$ has a cycle $(g+1, g+2, \ldots, w)$ such that $g=\operatorname{gcd}\left(w, s_{n}\right)$. By the observations in the proof of Claim 2, the numbers $g+1, \ldots, w-1$ all divide $s_{n}$; so the number $L=\operatorname{lcm}(g+1, g+2, \ldots, w-1)$ also divides $s_{n}$. Moreover, $g$ also divides $w$.

Now choose any $n \geqslant N$ such that $a_{n}=w$. By (1), we have

$$
s_{n+1}=g+\frac{w\left(s_{n}-w\right)}{g}=s_{n} \cdot \frac{w}{g}-\frac{w^{2}-g^{2}}{g} .
$$

Since $L$ divides both $s_{n}$ and $s_{n+1}$, it also divides the number $T=\frac{w^{2}-g^{2}}{g}$.
Suppose first that $w \geqslant 6$, which yields $g+1 \leqslant \frac{w}{2}+1 \leqslant w-2$. Then $(w-2)(w-1)|L| T$, so we have either $w^{2}-g^{2} \geqslant 2(w-1)(w-2)$, or $g=1$ and $w^{2}-g^{2}=(w-1)(w-2)$. In the former case we get $(w-1)(w-5)+\left(g^{2}-1\right) \leqslant 0$ which is false by our assumption. The latter equation rewrites as $3 w=3$, so $w=1$, which is also impossible.

Now we are left with the cases when $w \leqslant 5$ and $g \mid w$. The case $(w, g)=(4,1)$ violates the condition $L \left\lvert\, \frac{w^{2}-g^{2}}{g}\right.$; all other such pairs are listed in (2).

In the table below, for each pair $(w, g)$, we provide possible sequences $\left(a_{n}\right)$ and $\left(b_{n}\right)$. That shows that the cycles shown in (3) are indeed possible.

$$
\begin{array}{llll}
w=2 & g=1 & a_{n}=2 & b_{n}=2 \cdot 2^{n}+1 \\
w=3 & g=1 & \left(a_{2 k}, a_{2 k+1}\right)=(2,3) & \left(b_{2 k}, b_{2 k+1}\right)=\left(6 \cdot 3^{k}+2,6 \cdot 3^{k}+1\right) \\
w=4 & g=2 & \left(a_{2 k}, a_{2 k+1}\right)=(3,4) & \left(b_{2 k}, b_{2 k+1}\right)=\left(12 \cdot 2^{k}+3,12 \cdot 2^{k}+2\right) \\
w=5 & g=1 & \left(a_{4 k}, \ldots, a_{4 k+3}\right)=(2,3,4,5) & \left(b_{4 k}, \ldots, b_{4 k+3}\right)=\left(6 \cdot 5^{k}+4, \ldots, 6 \cdot 5^{k}+1\right)
\end{array}
$$

N5. Determine all triples $(a, b, c)$ of positive integers for which $a b-c, b c-a$, and $c a-b$ are powers of 2 .

Explanation: A power of 2 is an integer of the form $2^{n}$, where $n$ denotes some nonnegative integer.
(Serbia)
Answer. There are sixteen such triples, namely (2,2,2), the three permutations of $(2,2,3)$, and the six permutations of each of $(2,6,11)$ and $(3,5,7)$.
Solution 1. It can easily be verified that these sixteen triples are as required. Now let ( $a, b, c$ ) be any triple with the desired property. If we would have $a=1$, then both $b-c$ and $c-b$ were powers of 2 , which is impossible since their sum is zero; because of symmetry, this argument shows $a, b, c \geqslant 2$.

Case 1. Among $a, b$, and $c$ there are at least two equal numbers.
Without loss of generality we may suppose that $a=b$. Then $a^{2}-c$ and $a(c-1)$ are powers of 2 . The latter tells us that actually $a$ and $c-1$ are powers of 2 . So there are nonnegative integers $\alpha$ and $\gamma$ with $a=2^{\alpha}$ and $c=2^{\gamma}+1$. Since $a^{2}-c=2^{2 \alpha}-2^{\gamma}-1$ is a power of 2 and thus incongruent to -1 modulo 4 , we must have $\gamma \leqslant 1$. Moreover, each of the terms $2^{2 \alpha}-2$ and $2^{2 \alpha}-3$ can only be a power of 2 if $\alpha=1$. It follows that the triple $(a, b, c)$ is either $(2,2,2)$ or $(2,2,3)$.

Case 2. The numbers $a, b$, and $c$ are distinct.
Due to symmetry we may suppose that

$$
\begin{equation*}
2 \leqslant a<b<c . \tag{1}
\end{equation*}
$$

We are to prove that the triple $(a, b, c)$ is either $(2,6,11)$ or $(3,5,7)$. By our hypothesis, there exist three nonnegative integers $\alpha, \beta$, and $\gamma$ such that

$$
\begin{align*}
b c-a & =2^{\alpha},  \tag{2}\\
a c-b & =2^{\beta},  \tag{3}\\
\text { and } \quad a b-c & =2^{\gamma} . \tag{4}
\end{align*}
$$

Evidently we have

$$
\begin{equation*}
\alpha>\beta>\gamma . \tag{5}
\end{equation*}
$$

Depending on how large $a$ is, we divide the argument into two further cases.
Case 2.1. $\quad a=2$.
We first prove that $\gamma=0$. Assume for the sake of contradiction that $\gamma>0$. Then $c$ is even by (4) and, similarly, $b$ is even by (5) and (3). So the left-hand side of (2) is congruent to 2 modulo 4 , which is only possible if $b c=4$. As this contradicts (1), we have thereby shown that $\gamma=0$, i.e., that $c=2 b-1$.

Now (3) yields $3 b-2=2^{\beta}$. Due to $b>2$ this is only possible if $\beta \geqslant 4$. If $\beta=4$, then we get $b=6$ and $c=2 \cdot 6-1=11$, which is a solution. It remains to deal with the case $\beta \geqslant 5$. Now (2) implies

$$
9 \cdot 2^{\alpha}=9 b(2 b-1)-18=(3 b-2)(6 b+1)-16=2^{\beta}\left(2^{\beta+1}+5\right)-16,
$$

and by $\beta \geqslant 5$ the right-hand side is not divisible by 32 . Thus $\alpha \leqslant 4$ and we get a contradiction to (5).

Case 2.2. $\quad a \geqslant 3$.
Pick an integer $\vartheta \in\{-1,+1\}$ such that $c-\vartheta$ is not divisible by 4 . Now

$$
2^{\alpha}+\vartheta \cdot 2^{\beta}=\left(b c-a \vartheta^{2}\right)+\vartheta(c a-b)=(b+a \vartheta)(c-\vartheta)
$$

is divisible by $2^{\beta}$ and, consequently, $b+a \vartheta$ is divisible by $2^{\beta-1}$. On the other hand, $2^{\beta}=a c-b>$ $(a-1) c \geqslant 2 c$ implies in view of (1) that $a$ and $b$ are smaller than $2^{\beta-1}$. All this is only possible if $\vartheta=1$ and $a+b=2^{\beta-1}$. Now (3) yields

$$
\begin{equation*}
a c-b=2(a+b), \tag{6}
\end{equation*}
$$

whence $4 b>a+3 b=a(c-1) \geqslant a b$, which in turn yields $a=3$.
So (6) simplifies to $c=b+2$ and (2) tells us that $b(b+2)-3=(b-1)(b+3)$ is a power of 2 . Consequently, the factors $b-1$ and $b+3$ are powers of 2 themselves. Since their difference is 4 , this is only possible if $b=5$ and thus $c=7$. Thereby the solution is complete.

Solution 2. As in the beginning of the first solution, we observe that $a, b, c \geqslant 2$. Depending on the parities of $a, b$, and $c$ we distinguish three cases.

Case 1. The numbers $a, b$, and $c$ are even.
Let $2^{A}, 2^{B}$, and $2^{C}$ be the largest powers of 2 dividing $a, b$, and $c$ respectively. We may assume without loss of generality that $1 \leqslant A \leqslant B \leqslant C$. Now $2^{B}$ is the highest power of 2 dividing $a c-b$, whence $a c-b=2^{B} \leqslant b$. Similarly, we deduce $b c-a=2^{A} \leqslant a$. Adding both estimates we get $(a+b) c \leqslant 2(a+b)$, whence $c \leqslant 2$. So $c=2$ and thus $A=B=C=1$; moreover, we must have had equality throughout, i.e., $a=2^{A}=2$ and $b=2^{B}=2$. We have thereby found the solution $(a, b, c)=(2,2,2)$.

Case 2. The numbers $a, b$, and $c$ are odd.
If any two of these numbers are equal, say $a=b$, then $a c-b=a(c-1)$ has a nontrivial odd divisor and cannot be a power of 2 . Hence $a, b$, and $c$ are distinct. So we may assume without loss of generality that $a<b<c$.

Let $\alpha$ and $\beta$ denote the nonnegative integers for which $b c-a=2^{\alpha}$ and $a c-b=2^{\beta}$ hold. Clearly, we have $\alpha>\beta$, and thus $2^{\beta}$ divides

$$
a \cdot 2^{\alpha}-b \cdot 2^{\beta}=a(b c-a)-b(a c-b)=b^{2}-a^{2}=(b+a)(b-a) .
$$

Since $a$ is odd, it is not possible that both factors $b+a$ and $b-a$ are divisible by 4 . Consequently, one of them has to be a multiple of $2^{\beta-1}$. Hence one of the numbers $2(b+a)$ and $2(b-a)$ is divisible by $2^{\beta}$ and in either case we have

$$
\begin{equation*}
a c-b=2^{\beta} \leqslant 2(a+b) . \tag{7}
\end{equation*}
$$

This in turn yields $(a-1) b<a c-b<4 b$ and thus $a=3$ (recall that $a$ is odd and larger than 1). Substituting this back into (7) we learn $c \leqslant b+2$. But due to the parity $b<c$ entails that $b+2 \leqslant c$ holds as well. So we get $c=b+2$ and from $b c-a=(b-1)(b+3)$ being a power of 2 it follows that $b=5$ and $c=7$.

Case 3. Among $a, b$, and $c$ both parities occur.
Without loss of generality, we suppose that $c$ is odd and that $a \leqslant b$. We are to show that $(a, b, c)$ is either $(2,2,3)$ or $(2,6,11)$. As at least one of $a$ and $b$ is even, the expression $a b-c$ is odd; since it is also a power of 2 , we obtain

$$
\begin{equation*}
a b-c=1 . \tag{8}
\end{equation*}
$$

If $a=b$, then $c=a^{2}-1$, and from $a c-b=a\left(a^{2}-2\right)$ being a power of 2 it follows that both $a$ and $a^{2}-2$ are powers of 2 , whence $a=2$. This gives rise to the solution ( $2,2,3$ ).

We may suppose $a<b$ from now on. As usual, we let $\alpha>\beta$ denote the integers satisfying

$$
\begin{equation*}
2^{\alpha}=b c-a \quad \text { and } \quad 2^{\beta}=a c-b \tag{9}
\end{equation*}
$$

If $\beta=0$ it would follow that $a c-b=a b-c=1$ and hence that $b=c=1$, which is absurd. So $\beta$ and $\alpha$ are positive and consequently $a$ and $b$ are even. Substituting $c=a b-1$ into (9) we obtain

$$
\begin{align*}
2^{\alpha} & =a b^{2}-(a+b)  \tag{10}\\
\text { and } \quad 2^{\beta} & =a^{2} b-(a+b) . \tag{11}
\end{align*}
$$

The addition of both equation yields $2^{\alpha}+2^{\beta}=(a b-2)(a+b)$. Now $a b-2$ is even but not divisible by 4 , so the highest power of 2 dividing $a+b$ is $2^{\beta-1}$. For this reason, the equations (10) and (11) show that the highest powers of 2 dividing either of the numbers $a b^{2}$ and $a^{2} b$ is likewise $2^{\beta-1}$. Thus there is an integer $\tau \geqslant 1$ together with odd integers $A, B$, and $C$ such that $a=2^{\tau} A, b=2^{\tau} B, a+b=2^{3 \tau} C$, and $\beta=1+3 \tau$.

Notice that $A+B=2^{2 \tau} C \geqslant 4 C$. Moreover, (11) entails $A^{2} B-C=2$. Thus $8=$ $4 A^{2} B-4 C \geqslant 4 A^{2} B-A-B \geqslant A^{2}(3 B-1)$. Since $A$ and $B$ are odd with $A<B$, this is only possible if $A=1$ and $B=3$. Finally, one may conclude $C=1, \tau=1, a=2, b=6$, and $c=11$. We have thereby found the triple $(2,6,11)$. This completes the discussion of the third case, and hence the solution.

Comment. In both solutions, there are many alternative ways to proceed in each of its cases. Here we present a different treatment of the part " $a<b$ " of Case 3 in Solution 2, assuming that (8) and (9) have already been written down:

Put $d=\operatorname{gcd}(a, b)$ and define the integers $p$ and $q$ by $a=d p$ and $b=d q$; notice that $p<q$ and $\operatorname{gcd}(p, q)=1$. Now (8) implies $c=d^{2} p q-1$ and thus we have

$$
\begin{align*}
& 2^{\alpha}=d\left(d^{2} p q^{2}-p-q\right) \\
\text { and } \quad 2^{\beta} & =d\left(d^{2} p^{2} q-p-q\right) . \tag{12}
\end{align*}
$$

Now $2^{\beta}$ divides $2^{\alpha}-2^{\beta}=d^{3} p q(q-p)$ and, as $p$ and $q$ are easily seen to be coprime to $d^{2} p^{2} q-p-q$, it follows that

$$
\begin{equation*}
\left(d^{2} p^{2} q-p-q\right) \mid d^{2}(q-p) \tag{13}
\end{equation*}
$$

In particular, we have $d^{2} p^{2} q-p-q \leqslant d^{2}(q-p)$, i.e., $d^{2}\left(p^{2} q+p-q\right) \leqslant p+q$. As $p^{2} q+p-q>0$, this may be weakened to $p^{2} q+p-q \leqslant p+q$. Hence $p^{2} q \leqslant 2 q$, which is only possible if $p=1$.

Going back to (13), we get

$$
\begin{equation*}
\left(d^{2} q-q-1\right) \mid d^{2}(q-1) \tag{14}
\end{equation*}
$$

Now $2\left(d^{2} q-q-1\right) \leqslant d^{2}(q-1)$ would entail $d^{2}(q+1) \leqslant 2(q+1)$ and thus $d=1$. But this would tell us that $a=d p=1$, which is absurd. This argument proves $2\left(d^{2} q-q-1\right)>d^{2}(q-1)$ and in the light of (14) it follows that $d^{2} q-q-1=d^{2}(q-1)$, i.e., $q=d^{2}-1$. Plugging this together with $p=1$ into (12) we infer $2^{\beta}=d^{3}\left(d^{2}-2\right)$. Hence $d$ and $d^{2}-2$ are powers of 2 . Consequently, $d=2, q=3$, $a=2, b=6$, and $c=11$, as desired.

N6. Let $\mathbb{Z}_{>0}$ denote the set of positive integers. Consider a function $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$. For any $m, n \in \mathbb{Z}_{>0}$ we write $f^{n}(m)=\underbrace{f(f(\ldots f}_{n}(m) \ldots))$. Suppose that $f$ has the following two properties:
(i) If $m, n \in \mathbb{Z}_{>0}$, then $\frac{f^{n}(m)-m}{n} \in \mathbb{Z}_{>0}$;
(ii) The set $\mathbb{Z}_{>0} \backslash\left\{f(n) \mid n \in \mathbb{Z}_{>0}\right\}$ is finite.

Prove that the sequence $f(1)-1, f(2)-2, f(3)-3, \ldots$ is periodic.
(Singapore)
Solution. We split the solution into three steps. In the first of them, we show that the function $f$ is injective and explain how this leads to a useful visualization of $f$. Then comes the second step, in which most of the work happens: its goal is to show that for any $n \in \mathbb{Z}_{>0}$ the sequence $n, f(n), f^{2}(n), \ldots$ is an arithmetic progression. Finally, in the third step we put everything together, thus solving the problem.

Step 1. We commence by checking that $f$ is injective. For this purpose, we consider any $m, k \in \mathbb{Z}_{>0}$ with $f(m)=f(k)$. By $(i)$, every positive integer $n$ has the property that

$$
\frac{k-m}{n}=\frac{f^{n}(m)-m}{n}-\frac{f^{n}(k)-k}{n}
$$

is a difference of two integers and thus integral as well. But for $n=|k-m|+1$ this is only possible if $k=m$. Thereby, the injectivity of $f$ is established.

Now recall that due to condition (ii) there are finitely many positive integers $a_{1}, \ldots, a_{k}$ such that $\mathbb{Z}_{>0}$ is the disjoint union of $\left\{a_{1}, \ldots, a_{k}\right\}$ and $\left\{f(n) \mid n \in \mathbb{Z}_{>0}\right\}$. Notice that by plugging $n=1$ into condition $(i)$ we get $f(m)>m$ for all $m \in \mathbb{Z}_{>0}$.

We contend that every positive integer $n$ may be expressed uniquely in the form $n=f^{j}\left(a_{i}\right)$ for some $j \geqslant 0$ and $i \in\{1, \ldots, k\}$. The uniqueness follows from the injectivity of $f$. The existence can be proved by induction on $n$ in the following way. If $n \in\left\{a_{1}, \ldots, a_{k}\right\}$, then we may take $j=0$; otherwise there is some $n^{\prime}<n$ with $f\left(n^{\prime}\right)=n$ to which the induction hypothesis may be applied.

The result of the previous paragraph means that every positive integer appears exactly once in the following infinite picture, henceforth referred to as "the Table":

| $a_{1}$ | $f\left(a_{1}\right)$ | $f^{2}\left(a_{1}\right)$ | $f^{3}\left(a_{1}\right)$ | $\ldots$ |
| :---: | :---: | :---: | :---: | :---: |
| $a_{2}$ | $f\left(a_{2}\right)$ | $f^{2}\left(a_{2}\right)$ | $f^{3}\left(a_{2}\right)$ | $\ldots$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |  |
| $a_{k}$ | $f\left(a_{k}\right)$ | $f^{2}\left(a_{k}\right)$ | $f^{3}\left(a_{k}\right)$ | $\ldots$ |

The Table

Step 2. Our next goal is to prove that each row of the Table is an arithmetic progression. Assume contrariwise that the number $t$ of rows which are arithmetic progressions would satisfy $0 \leqslant t<k$. By permuting the rows if necessary we may suppose that precisely the first $t$ rows are arithmetic progressions, say with steps $T_{1}, \ldots, T_{t}$. Our plan is to find a further row that is "not too sparse" in an asymptotic sense, and then to prove that such a row has to be an arithmetic progression as well.

Let us write $T=\operatorname{lcm}\left(T_{1}, T_{2}, \ldots, T_{t}\right)$ and $A=\max \left\{a_{1}, a_{2}, \ldots, a_{t}\right\}$ if $t>0$; and $T=1$ and $A=0$ if $t=0$. For every integer $n \geqslant A$, the interval $\Delta_{n}=[n+1, n+T]$ contains exactly $T / T_{i}$
elements of the $i^{\text {th }}$ row $(1 \leqslant i \leqslant t)$. Therefore, the number of elements from the last $(k-t)$ rows of the Table contained in $\Delta_{n}$ does not depend on $n \geqslant A$. It is not possible that none of these intervals $\Delta_{n}$ contains an element from the $k-t$ last rows, because infinitely many numbers appear in these rows. It follows that for each $n \geqslant A$ the interval $\Delta_{n}$ contains at least one member from these rows.

This yields that for every positive integer $d$, the interval $[A+1, A+(d+1)(k-t) T]$ contains at least $(d+1)(k-t)$ elements from the last $k-t$ rows; therefore, there exists an index $x$ with $t+1 \leqslant x \leqslant k$, possibly depending on $d$, such that our interval contains at least $d+1$ elements from the $x^{\text {th }}$ row. In this situation we have

$$
f^{d}\left(a_{x}\right) \leqslant A+(d+1)(k-t) T
$$

Finally, since there are finitely many possibilities for $x$, there exists an index $x \geqslant t+1$ such that the set

$$
X=\left\{d \in \mathbb{Z}_{>0} \mid f^{d}\left(a_{x}\right) \leqslant A+(d+1)(k-t) T\right\}
$$

is infinite. Thereby we have found the "dense row" promised above.
By assumption ( $i$, for every $d \in X$ the number

$$
\beta_{d}=\frac{f^{d}\left(a_{x}\right)-a_{x}}{d}
$$

is a positive integer not exceeding

$$
\frac{A+(d+1)(k-t) T}{d} \leqslant \frac{A d+2 d(k-t) T}{d}=A+2(k-t) T
$$

This leaves us with finitely many choices for $\beta_{d}$, which means that there exists a number $T_{x}$ such that the set

$$
Y=\left\{d \in X \mid \beta_{d}=T_{x}\right\}
$$

is infinite. Notice that we have $f^{d}\left(a_{x}\right)=a_{x}+d \cdot T_{x}$ for all $d \in Y$.
Now we are prepared to prove that the numbers in the $x^{\text {th }}$ row form an arithmetic progression, thus coming to a contradiction with our assumption. Let us fix any positive integer $j$. Since the set $Y$ is infinite, we can choose a number $y \in Y$ such that $y-j>\left|f^{j}\left(a_{x}\right)-\left(a_{x}+j T_{x}\right)\right|$. Notice that both numbers

$$
f^{y}\left(a_{x}\right)-f^{j}\left(a_{x}\right)=f^{y-j}\left(f^{j}\left(a_{x}\right)\right)-f^{j}\left(a_{x}\right) \quad \text { and } \quad f^{y}\left(a_{x}\right)-\left(a_{x}+j T_{x}\right)=(y-j) T_{x}
$$

are divisible by $y-j$. Thus, the difference between these numbers is also divisible by $y-j$. Since the absolute value of this difference is less than $y-j$, it has to vanish, so we get $f^{j}\left(a_{x}\right)=$ $a_{x}+j \cdot T_{x}$.

Hence, it is indeed true that all rows of the Table are arithmetic progressions.
Step 3. Keeping the above notation in force, we denote the step of the $i^{\text {th }}$ row of the table by $T_{i}$. $\overline{\text { Now we claim that we have } f(n)-n=f(n+T)-(n+T) \text { for all } n \in \mathbb{Z}_{>0} \text {, where }{ }^{\prime} \text {, } n(n)}$

$$
T=\operatorname{lcm}\left(T_{1}, \ldots, T_{k}\right) .
$$

To see this, let any $n \in \mathbb{Z}_{>0}$ be given and denote the index of the row in which it appears in the Table by $i$. Then we have $f^{j}(n)=n+j \cdot T_{i}$ for all $j \in \mathbb{Z}_{>0}$, and thus indeed

$$
f(n+T)-f(n)=f^{1+T / T_{i}}(n)-f(n)=\left(n+T+T_{i}\right)-\left(n+T_{i}\right)=T
$$

This concludes the solution.

Comment 1. There are some alternative ways to complete the second part once the index $x$ corresponding to a "dense row" is found. For instance, one may show that for some integer $T_{x}^{*}$ the set

$$
Y^{*}=\left\{j \in \mathbb{Z}_{>0} \mid f^{j+1}\left(a_{x}\right)-f^{j}\left(a_{x}\right)=T_{x}^{*}\right\}
$$

is infinite, and then one may conclude with a similar divisibility argument.
Comment 2. It may be checked that, conversely, any way to fill out the Table with finitely many arithmetic progressions so that each positive integer appears exactly once, gives rise to a function $f$ satisfying the two conditions mentioned in the problem. For example, we may arrange the positive integers as follows:

| 2 | 4 | 6 | 8 | 10 | $\ldots$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 5 | 9 | 13 | 17 | $\ldots$ |
| 3 | 7 | 11 | 15 | 19 | $\ldots$ |

This corresponds to the function

$$
f(n)= \begin{cases}n+2 & \text { if } n \text { is even; } \\ n+4 & \text { if } n \text { is odd }\end{cases}
$$

As this example shows, it is not true that the function $n \mapsto f(n)-n$ has to be constant.

N7. Let $\mathbb{Z}_{>0}$ denote the set of positive integers. For any positive integer $k$, a function $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ is called $k$-good if $\operatorname{gcd}(f(m)+n, f(n)+m) \leqslant k$ for all $m \neq n$. Find all $k$ such that there exists a $k$-good function.
(Canada)
Answer. $k \geqslant 2$.
Solution 1. For any function $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$, let $G_{f}(m, n)=\operatorname{gcd}(f(m)+n, f(n)+m)$. Note that a $k$-good function is also $(k+1)$-good for any positive integer $k$. Hence, it suffices to show that there does not exist a 1-good function and that there exists a 2 -good function.

We first show that there is no 1 -good function. Suppose that there exists a function $f$ such that $G_{f}(m, n)=1$ for all $m \neq n$. Now, if there are two distinct even numbers $m$ and $n$ such that $f(m)$ and $f(n)$ are both even, then $2 \mid G_{f}(m, n)$, a contradiction. A similar argument holds if there are two distinct odd numbers $m$ and $n$ such that $f(m)$ and $f(n)$ are both odd. Hence we can choose an even $m$ and an odd $n$ such that $f(m)$ is odd and $f(n)$ is even. This also implies that $2 \mid G_{f}(m, n)$, a contradiction.

We now construct a 2 -good function. Define $f(n)=2^{g(n)+1}-n-1$, where $g$ is defined recursively by $g(1)=1$ and $g(n+1)=\left(2^{g(n)+1}\right)$ !.

For any positive integers $m>n$, set

$$
A=f(m)+n=2^{g(m)+1}-m+n-1, \quad B=f(n)+m=2^{g(n)+1}-n+m-1
$$

We need to show that $\operatorname{gcd}(A, B) \leqslant 2$. First, note that $A+B=2^{g(m)+1}+2^{g(n)+1}-2$ is not divisible by 4 , so that $4 \nmid \operatorname{gcd}(A, B)$. Now we suppose that there is an odd prime $p$ for which $p \mid \operatorname{gcd}(A, B)$ and derive a contradiction.

We first claim that $2^{g(m-1)+1} \geqslant B$. This is a rather weak bound; one way to prove it is as follows. Observe that $g(k+1)>g(k)$ and hence $2^{g(k+1)+1} \geqslant 2^{g(k)+1}+1$ for every positive integer $k$. By repeatedly applying this inequality, we obtain $2^{g(m-1)+1} \geqslant 2^{g(n)+1}+(m-1)-n=B$.

Now, since $p \mid B$, we have $p-1<B \leqslant 2^{g(m-1)+1}$, so that $p-1 \mid\left(2^{g(m-1)+1}\right)!=g(m)$. Hence $2^{g(m)} \equiv 1(\bmod p)$, which yields $A+B \equiv 2^{g(n)+1}(\bmod p)$. However, since $p \mid A+B$, this implies that $p=2$, a contradiction.

Solution 2. We provide an alternative construction of a 2 -good function $f$.
Let $\mathcal{P}$ be the set consisting of 4 and all odd primes. For every $p \in \mathcal{P}$, we say that a number $a \in\{0,1, \ldots, p-1\}$ is $p$-useful if $a \not \equiv-a(\bmod p)$. Note that a residue modulo $p$ which is neither 0 nor 2 is $p$-useful (the latter is needed only when $p=4$ ).

We will construct $f$ recursively; in some steps, we will also define a $p$-useful number $a_{p}$. After the $m^{\text {th }}$ step, the construction will satisfy the following conditions:
( $i$ ) The values of $f(n)$ have already been defined for all $n \leqslant m$, and $p$-useful numbers $a_{p}$ have already been defined for all $p \leqslant m+2$;
(ii) If $n \leqslant m$ and $p \leqslant m+2$, then $f(n)+n \not \equiv a_{p}(\bmod p)$;
(iii) $\operatorname{gcd}\left(f\left(n_{1}\right)+n_{2}, f\left(n_{2}\right)+n_{1}\right) \leqslant 2$ for all $n_{1}<n_{2} \leqslant m$.

If these conditions are satisfied, then $f$ will be a 2 -good function.
Step 1. Set $f(1)=1$ and $a_{3}=1$. Clearly, all the conditions are satisfied.
Step $m$, for $m \geqslant 2$. We need to determine $f(m)$ and, if $m+2 \in \mathcal{P}$, the number $a_{m+2}$. Defining $f(m)$. Let $X_{m}=\{p \in \mathcal{P}: p \mid f(n)+m$ for some $n<m\}$. We will determine $f(m) \bmod p$ for all $p \in X_{m}$ and then choose $f(m)$ using the Chinese Remainder Theorem.

Take any $p \in X_{m}$. If $p \leqslant m+1$, then we define $f(m) \equiv-a_{p}-m(\bmod p)$. Otherwise, if $p \geqslant m+2$, then we define $f(m) \equiv 0(\bmod p)$.
Defining $a_{m+2}$. Now let $p=m+2$ and suppose that $p \in \mathcal{P}$. We choose $a_{p}$ to be a residue modulo $p$ that is not congruent to 0,2 , or $f(n)+n$ for any $n \leqslant m$. Since $f(1)+1=2$, there are at most $m+1<p$ residues to avoid, so we can always choose a remaining residue.

We first check that ( $i i$ ) is satisfied. We only need to check it if $p=m+2$ or $n=m$. In the former case, we have $f(n)+n \not \equiv a_{p}(\bmod p)$ by construction. In the latter case, if $n=m$ and $p \leqslant m+1$, then we have $f(m)+m \equiv-a_{p} \not \equiv a_{p}(\bmod p)$, where we make use of the fact that $a_{p}$ is $p$-useful.

Now we check that (iii) holds. Suppose, to the contrary, that $p \mid \operatorname{gcd}(f(n)+m, f(m)+n)$ for some $n<m$. Then $p \in X_{m}$ and $p \mid f(m)+n$. If $p \geqslant m+2$, then $0 \equiv f(m)+n \equiv n(\bmod p)$, which is impossible since $n<m<p$.

Otherwise, if $p \leqslant m+1$, then

$$
0 \equiv(f(m)+n)+(f(n)+m) \equiv(f(n)+n)+(f(m)+m) \equiv(f(n)+n)-a_{p} \quad(\bmod p) .
$$

This implies that $f(n)+n \equiv a_{p}(\bmod p)$, a contradiction with $(i i)$.
Comment 1. For any $p \in \mathcal{P}$, we may also define $a_{p}$ at step $m$ for an arbitrary $m \leqslant p-2$. The construction will work as long as we define a finite number of $a_{p}$ at each step.

Comment 2. When attempting to construct a 2 -good function $f$ recursively, the following way seems natural. Start with setting $f(1)=1$. Next, for each integer $m>1$, introduce the set $X_{m}$ like in Solution 2 and define $f(m)$ so as to satisfy

$$
\begin{array}{rll}
f(m) \equiv f(m-p) & (\bmod p) & \text { for all } p \in X_{m} \text { with } p<m, \quad \text { and } \\
f(m) \equiv 0 & (\bmod p) & \text { for all } p \in X_{m} \text { with } p \geqslant m .
\end{array}
$$

This construction might seem to work. Indeed, consider a fixed $p \in \mathcal{P}$, and suppose that $p$ divides $\operatorname{gcd}(f(n)+m, f(m)+n)$ for some $n<m$. Choose such $m$ and $n$ so that $\max (m, n)$ is minimal. Then $p \in X_{m}$. We can check that $p<m$, so that the construction implies that $p$ divides $\operatorname{gcd}(f(n)+(m-p), f(m-p)+n)$. Since $\max (n, m-p)<\max (m, n)$, this almost leads to a contradiction - the only trouble is the possibility that $n=m-p$. However, this flaw may happen to be not so easy to fix.

We will present one possible way to repair this argument in the next comment.
Comment 3. There are many recursive constructions for a 2 -good function $f$. Here we sketch one general approach which may be specified in different ways. For convenience, we denote by $\mathbb{Z}_{p}$ the set of residues modulo $p$; all operations on elements of $\mathbb{Z}_{p}$ are also performed modulo $p$.

The general structure is the same as in Solution 2, i.e. using the Chinese Remainder Theorem to successively determine $f(m)$. But instead of designating a common "safe" residue $a_{p}$ for future steps, we act as follows.

For every $p \in \mathcal{P}$, in some step of the process we define $p$ subsets $B_{p}^{(1)}, B_{p}^{(2)}, \ldots, B_{p}^{(p)} \subset \mathbb{Z}_{p}$. The meaning of these sets is that

$$
\begin{equation*}
f(m)+m \text { should be congruent to some element in } B_{p}^{(i)} \text { whenever } m \equiv i(\bmod p) \text { for } i \in \mathbb{Z}_{p} \tag{1}
\end{equation*}
$$

Moreover, in every such subset we specify a safe element $b_{p}^{(i)} \in B_{p}^{(i)}$. The meaning now is that in future steps, it is safe to set $f(m)+m \equiv b_{p}^{(i)}(\bmod p)$ whenever $m \equiv i(\bmod p)$. In view of (1), this safety will follow from the condition that $p \nmid \operatorname{gcd}\left(b_{p}^{(i)}+(j-i), c^{(j)}-(j-i)\right)$ for all $j \in \mathbb{Z}_{p}$ and all $c^{(j)} \in B_{p}^{(j)}$. In turn, this condition can be rewritten as

$$
\begin{equation*}
-b_{p}^{(i)} \notin B_{p}^{(j)}, \quad \text { where } \quad j \equiv i-b_{p}^{(i)} \quad(\bmod p) . \tag{2}
\end{equation*}
$$

The construction in Solution 2 is equivalent to setting $b_{p}^{(i)}=-a_{p}$ and $B_{p}^{(i)}=\mathbb{Z}_{p} \backslash\left\{a_{p}\right\}$ for all $i$. However, there are different, more technical specifications of our approach.

One may view the (incomplete) construction in Comment 2 as defining $B_{p}^{(i)}$ and $b_{p}^{(i)}$ at step $p-1$ by setting $B_{p}^{(0)}=\left\{b_{p}^{(0)}\right\}=\{0\}$ and $B_{p}^{(i)}=\left\{b_{p}^{(i)}\right\}=\{f(i)+i \bmod p\}$ for every $i=1,2, \ldots, p-1$. However, this construction violates (2) as soon as some number of the form $f(i)+i$ is divisible by some $p$ with $i+2 \leqslant p \in \mathcal{P}$, since then $-b_{p}^{(i)}=b_{p}^{(i)} \in B_{p}^{(i)}$.

Here is one possible way to repair this construction. For all $p \in \mathcal{P}$, we define the sets $B_{p}^{(i)}$ and the elements $b_{p}^{(i)}$ at step $(p-2)$ as follows. Set $B_{p}^{(1)}=\left\{b_{p}^{(1)}\right\}=\{2\}$ and $B_{p}^{(-1)}=B_{p}^{(0)}=\left\{b_{p}^{(-1)}\right\}=\left\{b_{p}^{(0)}\right\}=$ $\{-1\}$. Next, for all $i=2, \ldots, p-2$, define $B_{p}^{(i)}=\{i, f(i)+i \bmod p\}$ and $b_{p}^{(i)}=i$. One may see that these definitions agree with both (1) and (2).

N8. For every positive integer $n$ with prime factorization $n=\prod_{i=1}^{k} p_{i}^{\alpha_{i}}$, define

$$
\mathcal{V}(n)=\sum_{i: p_{i}>10^{100}} \alpha_{i} .
$$

That is, $\mho(n)$ is the number of prime factors of $n$ greater than $10^{100}$, counted with multiplicity.
Find all strictly increasing functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that

$$
\begin{equation*}
\mho(f(a)-f(b)) \leqslant \mho(a-b) \quad \text { for all integers } a \text { and } b \text { with } a>b . \tag{1}
\end{equation*}
$$

(Brazil)
Answer. $f(x)=a x+b$, where $b$ is an arbitrary integer, and $a$ is an arbitrary positive integer with $\mho(a)=0$.
Solution. A straightforward check shows that all the functions listed in the answer satisfy the problem condition. It remains to show the converse.

Assume that $f$ is a function satisfying the problem condition. Notice that the function $g(x)=f(x)-f(0)$ also satisfies this condition. Replacing $f$ by $g$, we assume from now on that $f(0)=0$; then $f(n)>0$ for any positive integer $n$. Thus, we aim to prove that there exists a positive integer $a$ with $\mho(a)=0$ such that $f(n)=a n$ for all $n \in \mathbb{Z}$.

We start by introducing some notation. Set $N=10^{100}$. We say that a prime $p$ is large if $p>N$, and $p$ is small otherwise; let $\mathcal{S}$ be the set of all small primes. Next, we say that a positive integer is large or small if all its prime factors are such (thus, the number 1 is the unique number which is both large and small). For a positive integer $k$, we denote the greatest large divisor of $k$ and the greatest small divisor of $k$ by $L(k)$ and $S(k)$, respectively; thus, $k=L(k) S(k)$.

We split the proof into three steps.
Step 1. We prove that for every large $k$, we have $k|f(a)-f(b) \Longleftrightarrow k| a-b$. In other words, $L(f(a)-f(b))=L(a-b)$ for all integers $a$ and $b$ with $a>b$.

We use induction on $k$. The base case $k=1$ is trivial. For the induction step, assume that $k_{0}$ is a large number, and that the statement holds for all large numbers $k$ with $k<k_{0}$.
Claim 1. For any integers $x$ and $y$ with $0<x-y<k_{0}$, the number $k_{0}$ does not divide $f(x)-f(y)$.
Proof. Assume, to the contrary, that $k_{0} \mid f(x)-f(y)$. Let $\ell=L(x-y)$; then $\ell \leqslant x-y<k_{0}$. By the induction hypothesis, $\ell \mid f(x)-f(y)$, and thus $\operatorname{lcm}\left(k_{0}, \ell\right) \mid f(x)-f(y)$. Notice that $\operatorname{lcm}\left(k_{0}, \ell\right)$ is large, and $\operatorname{lcm}\left(k_{0}, \ell\right) \geqslant k_{0}>\ell$. But then

$$
\mho(f(x)-f(y)) \geqslant \mho\left(\operatorname{lcm}\left(k_{0}, \ell\right)\right)>\mho(\ell)=\mho(x-y),
$$

which is impossible.
Now we complete the induction step. By Claim 1, for every integer $a$ each of the sequences

$$
f(a), f(a+1), \ldots, f\left(a+k_{0}-1\right) \quad \text { and } \quad f(a+1), f(a+2), \ldots, f\left(a+k_{0}\right)
$$

forms a complete residue system modulo $k_{0}$. This yields $f(a) \equiv f\left(a+k_{0}\right)\left(\bmod k_{0}\right)$. Thus, $f(a) \equiv f(b)\left(\bmod k_{0}\right)$ whenever $a \equiv b\left(\bmod k_{0}\right)$.

Finally, if $a \not \equiv b\left(\bmod k_{0}\right)$ then there exists an integer $b^{\prime}$ such that $b^{\prime} \equiv b\left(\bmod k_{0}\right)$ and $\left|a-b^{\prime}\right|<k_{0}$. Then $f(b) \equiv f\left(b^{\prime}\right) \not \equiv f(a)\left(\bmod k_{0}\right)$. The induction step is proved.
Step 2. We prove that for some small integer a there exist infinitely many integers $n$ such that $\overline{f(n)=}$ an. In other words, $f$ is linear on some infinite set.

We start with the following general statement.

Claim 2. There exists a constant $c$ such that $f(t)<c t$ for every positive integer $t>N$.
Proof. Let $d$ be the product of all small primes, and let $\alpha$ be a positive integer such that $2^{\alpha}>f(N)$. Then, for every $p \in \mathcal{S}$ the numbers $f(0), f(1), \ldots, f(N)$ are distinct modulo $p^{\alpha}$. Set $P=d^{\alpha}$ and $c=P+f(N)$.

Choose any integer $t>N$. Due to the choice of $\alpha$, for every $p \in \mathcal{S}$ there exists at most one nonnegative integer $i \leqslant N$ with $p^{\alpha} \mid f(t)-f(i)$. Since $|\mathcal{S}|<N$, we can choose a nonnegative integer $j \leqslant N$ such that $p^{\alpha} \nmid f(t)-f(j)$ for all $p \in \mathcal{S}$. Therefore, $S(f(t)-f(j))<P$.

On the other hand, Step 1 shows that $L(f(t)-f(j))=L(t-j) \leqslant t-j$. Since $0 \leqslant j \leqslant N$, this yields

$$
f(t)=f(j)+L(f(t)-f(j)) \cdot S(f(t)-f(j))<f(N)+(t-j) P \leqslant(P+f(N)) t=c t .
$$

Now let $\mathcal{T}$ be the set of large primes. For every $t \in \mathcal{T}$, Step 1 implies $L(f(t))=t$, so the ratio $f(t) / t$ is an integer. Now Claim 2 leaves us with only finitely many choices for this ratio, which means that there exists an infinite subset $\mathcal{T}^{\prime} \subseteq \mathcal{T}$ and a positive integer $a$ such that $f(t)=a t$ for all $t \in \mathcal{T}^{\prime}$, as required.

Since $L(t)=L(f(t))=L(a) L(t)$ for all $t \in \mathcal{T}^{\prime}$, we get $L(a)=1$, so the number $a$ is small. Step 3. We show that $f(x)=$ ax for all $x \in \mathbb{Z}$.

Let $R_{i}=\{x \in \mathbb{Z}: x \equiv i(\bmod N!)\}$ denote the residue class of $i$ modulo $N!$.
Claim 3. Assume that for some $r$, there are infinitely many $n \in R_{r}$ such that $f(n)=a n$. Then $f(x)=a x$ for all $x \in R_{r+1}$.
Proof. Choose any $x \in R_{r+1}$. By our assumption, we can select $n \in R_{r}$ such that $f(n)=a n$ and $|n-x|>|f(x)-a x|$. Since $n-x \equiv r-(r+1)=-1(\bmod N!)$, the number $|n-x|$ is large. Therefore, by Step 1 we have $f(x) \equiv f(n)=a n \equiv a x(\bmod n-x)$, so $n-x \mid f(x)-a x$. Due to the choice of $n$, this yields $f(x)=a x$.

To complete Step 3, notice that the set $\mathcal{T}^{\prime}$ found in Step 2 contains infinitely many elements of some residue class $R_{i}$. Applying Claim 3, we successively obtain that $f(x)=a x$ for all $x \in R_{i+1}, R_{i+2}, \ldots, R_{i+N!}=R_{i}$. This finishes the solution.

Comment 1. As the proposer also mentions, one may also consider the version of the problem where the condition (1) is replaced by the condition that $L(f(a)-f(b))=L(a-b)$ for all integers $a$ and $b$ with $a>b$. This allows to remove of Step 1 from the solution.

Comment 2. Step 2 is the main step of the solution. We sketch several different approaches allowing to perform this step using statements which are weaker than Claim 2.
Approach 1. Let us again denote the product of all small primes by $d$. We focus on the values $f\left(d^{i}\right)$, $i \geqslant 0$. In view of Step 1, we have $L\left(f\left(d^{i}\right)-f\left(d^{k}\right)\right)=L\left(d^{i}-d^{k}\right)=d^{i-k}-1$ for all $i>k \geqslant 0$.

Acting similarly to the beginning of the proof of Claim 2, one may choose a number $\alpha \geqslant 0$ such that the residues of the numbers $f\left(d^{i}\right), i=0,1, \ldots, N$, are distinct modulo $p^{\alpha}$ for each $p \in \mathcal{S}$. Then, for every $i>N$, there exists an exponent $k=k(i) \leqslant N$ such that $S\left(f\left(d^{i}\right)-f\left(d^{k}\right)\right)<P=d^{\alpha}$.

Since there are only finitely many options for $k(i)$, as well as for the corresponding numbers $S\left(f\left(d^{i}\right)-f\left(d^{k}\right)\right)$, there exists an infinite set $I$ of exponents $i>N$ such that $k(i)$ attains the same value $k_{0}$ for all $i \in I$, and such that, moreover, $S\left(f\left(d^{i}\right)-f\left(d^{k_{0}}\right)\right)$ attains the same value $s_{0}$ for all $i \in I$. Therefore, for all such $i$ we have

$$
f\left(d^{i}\right)=f\left(d^{k_{0}}\right)+L\left(f\left(d^{i}\right)-f\left(d^{k_{0}}\right)\right) \cdot S\left(f\left(d^{i}\right)-f\left(d^{k_{0}}\right)\right)=f\left(d^{k_{0}}\right)+\left(d^{i-k_{0}}-1\right) s_{0},
$$

which means that $f$ is linear on the infinite set $\left\{d^{i}: i \in I\right\}$ (although with rational coefficients).
Finally, one may implement the relation $f\left(d^{i}\right) \equiv f(1)\left(\bmod d^{i}-1\right)$ in order to establish that in fact $f\left(d^{i}\right) / d^{i}$ is a (small and fixed) integer for all $i \in I$.

Approach 2. Alternatively, one may start with the following lemma.
Lemma. There exists a positive constant $c$ such that

$$
L\left(\prod_{i=1}^{3 N}(f(k)-f(i))\right)=\prod_{i=1}^{3 N} L(f(k)-f(i)) \geqslant c(f(k))^{2 N}
$$

for all $k>3 N$.
Proof. Let $k$ be an integer with $k>3 N$. Set $\Pi=\prod_{i=1}^{3 N}(f(k)-f(i))$.
Notice that for every prime $p \in \mathcal{S}$, at most one of the numbers in the set

$$
\mathcal{H}=\{f(k)-f(i): 1 \leqslant i \leqslant 3 N\}
$$

is divisible by a power of $p$ which is greater than $f(3 N)$; we say that such elements of $\mathcal{H}$ are bad. Now, for each element $h \in \mathcal{H}$ which is not bad we have $S(h) \leqslant f(3 N)^{N}$, while the bad elements do not exceed $f(k)$. Moreover, there are less than $N$ bad elements in $\mathcal{H}$. Therefore,

$$
S(\Pi)=\prod_{h \in \mathcal{H}} S(h) \leqslant(f(3 N))^{3 N^{2}} \cdot(f(k))^{N} .
$$

This easily yields the lemma statement in view of the fact that $L(\Pi) S(\Pi)=\Pi \geqslant \mu(f(k))^{3 N}$ for some absolute constant $\mu$.

As a corollary of the lemma, one may get a weaker version of Claim 2 stating that there exists a positive constant $C$ such that $f(k) \leqslant C k^{3 / 2}$ for all $k>3 N$. Indeed, from Step 1 we have

$$
k^{3 N} \geqslant \prod_{i=1}^{3 N} L(k-i)=\prod_{i=1}^{3 N} L(f(k)-f(i)) \geqslant c(f(k))^{2 N},
$$

so $f(k) \leqslant c^{-1 /(2 N)} k^{3 / 2}$.
To complete Step 2 now, set $a=f(1)$. Due to the estimates above, we may choose a positive integer $n_{0}$ such that $|f(n)-a n|<\frac{n(n-1)}{2}$ for all $n \geqslant n_{0}$.

Take any $n \geqslant n_{0}$ with $n \equiv 2(\bmod N!)$. Then $L(f(n)-f(0))=L(n)=n / 2$ and $L(f(n)-f(1))=$ $L(n-1)=n-1$; these relations yield $f(n) \equiv f(0)=0 \equiv a n(\bmod n / 2)$ and $f(n) \equiv f(1)=a \equiv a n$ $(\bmod n-1)$, respectively. Thus, $\left.\frac{n(n-1)}{2} \right\rvert\, f(n)-a n$, which shows that $f(n)=a n$ in view of the estimate above.

Comment 3. In order to perform Step 3, it suffices to establish the equality $f(n)=a n$ for any infinite set of values of $n$. However, if this set has some good structure, then one may find easier ways to complete this step.

For instance, after showing, as in Approach 2, that $f(n)=a n$ for all $n \geqslant n_{0}$ with $n \equiv 2(\bmod N!)$, one may proceed as follows. Pick an arbitrary integer $x$ and take any large prime $p$ which is greater than $|f(x)-a x|$. By the Chinese Remainder Theorem, there exists a positive integer $n>\max \left(x, n_{0}\right)$ such that $n \equiv 2(\bmod N!)$ and $n \equiv x(\bmod p)$. By Step 1 , we have $f(x) \equiv f(n)=a n \equiv a x(\bmod p)$. Due to the choice of $p$, this is possible only if $f(x)=a x$.

## CHIANG MAI, THAILAND 4-16 JULY 2015

# Shortlisted Problems with Solutions $57^{\text {th }}$ International Mathematical Olympiad Hong Kong, 2016 

## Note of Confidentiality

## The shortlisted problems should be kept strictly confidential until IMO 2017.

## Contributing Countries

The Organising Committee and the Problem Selection Committee of IMO 2016 thank the following 40 countries for contributing 121 problem proposals:

Albania, Algeria, Armenia, Australia, Austria, Belarus, Belgium, Bulgaria, Colombia, Cyprus, Czech Republic, Denmark, Estonia, France, Georgia, Greece, Iceland, India, Iran, Ireland, Israel, Japan, Latvia, Luxembourg, Malaysia, Mexico, Mongolia, Netherlands, Philippines, Russia, Serbia, Slovakia, Slovenia, South Africa, Taiwan, Tanzania, Thailand, Trinidad and Tobago, Turkey, Ukraine.

## Problem Selection Committee



Front row from left: Yong-Gao Chen, Andy Liu, Tat Wing Leung (Chairman).
Back row from left: Yi-Jun Yao, Yun-Hao Fu, Yi-Jie He, Zhongtao Wu, Heung Wing Joseph Lee, Chi Hong Chow, Ka Ho Law, Tak Wing Ching.

## Problems

## Algebra

A1. Let $a, b$ and $c$ be positive real numbers such that $\min \{a b, b c, c a\} \geqslant 1$. Prove that

$$
\sqrt[3]{\left(a^{2}+1\right)\left(b^{2}+1\right)\left(c^{2}+1\right)} \leqslant\left(\frac{a+b+c}{3}\right)^{2}+1
$$

A2. Find the smallest real constant $C$ such that for any positive real numbers $a_{1}, a_{2}, a_{3}, a_{4}$ and $a_{5}$ (not necessarily distinct), one can always choose distinct subscripts $i, j, k$ and $l$ such that

$$
\left|\frac{a_{i}}{a_{j}}-\frac{a_{k}}{a_{l}}\right| \leqslant C
$$

A3. Find all integers $n \geqslant 3$ with the following property: for all real numbers $a_{1}, a_{2}, \ldots, a_{n}$ and $b_{1}, b_{2}, \ldots, b_{n}$ satisfying $\left|a_{k}\right|+\left|b_{k}\right|=1$ for $1 \leqslant k \leqslant n$, there exist $x_{1}, x_{2}, \ldots, x_{n}$, each of which is either -1 or 1 , such that

$$
\left|\sum_{k=1}^{n} x_{k} a_{k}\right|+\left|\sum_{k=1}^{n} x_{k} b_{k}\right| \leqslant 1
$$

A4. Denote by $\mathbb{R}^{+}$the set of all positive real numbers. Find all functions $f: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$such that

$$
x f\left(x^{2}\right) f(f(y))+f(y f(x))=f(x y)\left(f\left(f\left(x^{2}\right)\right)+f\left(f\left(y^{2}\right)\right)\right)
$$

for all positive real numbers $x$ and $y$.

## A5.

(a) Prove that for every positive integer $n$, there exists a fraction $\frac{a}{b}$ where $a$ and $b$ are integers satisfying $0<b \leqslant \sqrt{n}+1$ and $\sqrt{n} \leqslant \frac{a}{b} \leqslant \sqrt{n+1}$.
(b) Prove that there are infinitely many positive integers $n$ such that there is no fraction $\frac{a}{b}$ where $a$ and $b$ are integers satisfying $0<b \leqslant \sqrt{n}$ and $\sqrt{n} \leqslant \frac{a}{b} \leqslant \sqrt{n+1}$.

A6. The equation

$$
(x-1)(x-2) \cdots(x-2016)=(x-1)(x-2) \cdots(x-2016)
$$

is written on the board. One tries to erase some linear factors from both sides so that each side still has at least one factor, and the resulting equation has no real roots. Find the least number of linear factors one needs to erase to achieve this.

A7. Denote by $\mathbb{R}$ the set of all real numbers. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ such that $f(0) \neq 0$ and

$$
f(x+y)^{2}=2 f(x) f(y)+\max \left\{f\left(x^{2}\right)+f\left(y^{2}\right), f\left(x^{2}+y^{2}\right)\right\}
$$

for all real numbers $x$ and $y$.

A8. Determine the largest real number $a$ such that for all $n \geqslant 1$ and for all real numbers $x_{0}, x_{1}, \ldots, x_{n}$ satisfying $0=x_{0}<x_{1}<x_{2}<\cdots<x_{n}$, we have

$$
\frac{1}{x_{1}-x_{0}}+\frac{1}{x_{2}-x_{1}}+\cdots+\frac{1}{x_{n}-x_{n-1}} \geqslant a\left(\frac{2}{x_{1}}+\frac{3}{x_{2}}+\cdots+\frac{n+1}{x_{n}}\right)
$$

## Combinatorics

C1. The leader of an IMO team chooses positive integers $n$ and $k$ with $n>k$, and announces them to the deputy leader and a contestant. The leader then secretly tells the deputy leader an $n$-digit binary string, and the deputy leader writes down all $n$-digit binary strings which differ from the leader's in exactly $k$ positions. (For example, if $n=3$ and $k=1$, and if the leader chooses 101, the deputy leader would write down 001, 111 and 100.) The contestant is allowed to look at the strings written by the deputy leader and guess the leader's string. What is the minimum number of guesses (in terms of $n$ and $k$ ) needed to guarantee the correct answer?

C2. Find all positive integers $n$ for which all positive divisors of $n$ can be put into the cells of a rectangular table under the following constraints:

- each cell contains a distinct divisor;
- the sums of all rows are equal; and
- the sums of all columns are equal.

C3. Let $n$ be a positive integer relatively prime to 6 . We paint the vertices of a regular $n$-gon with three colours so that there is an odd number of vertices of each colour. Show that there exists an isosceles triangle whose three vertices are of different colours.

C4. Find all positive integers $n$ for which we can fill in the entries of an $n \times n$ table with the following properties:

- each entry can be one of $I, M$ and $O$;
- in each row and each column, the letters $I, M$ and $O$ occur the same number of times; and
- in any diagonal whose number of entries is a multiple of three, the letters $I, M$ and $O$ occur the same number of times.

C5. Let $n \geqslant 3$ be a positive integer. Find the maximum number of diagonals of a regular $n$-gon one can select, so that any two of them do not intersect in the interior or they are perpendicular to each other.

C6. There are $n \geqslant 3$ islands in a city. Initially, the ferry company offers some routes between some pairs of islands so that it is impossible to divide the islands into two groups such that no two islands in different groups are connected by a ferry route.

After each year, the ferry company will close a ferry route between some two islands $X$ and $Y$. At the same time, in order to maintain its service, the company will open new routes according to the following rule: for any island which is connected by a ferry route to exactly one of $X$ and $Y$, a new route between this island and the other of $X$ and $Y$ is added.

Suppose at any moment, if we partition all islands into two nonempty groups in any way, then it is known that the ferry company will close a certain route connecting two islands from the two groups after some years. Prove that after some years there will be an island which is connected to all other islands by ferry routes.
$\mathbf{C 7}$. Let $n \geqslant 2$ be an integer. In the plane, there are $n$ segments given in such a way that any two segments have an intersection point in the interior, and no three segments intersect at a single point. Jeff places a snail at one of the endpoints of each of the segments and claps his hands $n-1$ times. Each time when he claps his hands, all the snails move along their own segments and stay at the next intersection points until the next clap. Since there are $n-1$ intersection points on each segment, all snails will reach the furthest intersection points from their starting points after $n-1$ claps.
(a) Prove that if $n$ is odd then Jeff can always place the snails so that no two of them ever occupy the same intersection point.
(b) Prove that if $n$ is even then there must be a moment when some two snails occupy the same intersection point no matter how Jeff places the snails.

C8. Let $n$ be a positive integer. Determine the smallest positive integer $k$ with the following property: it is possible to mark $k$ cells on a $2 n \times 2 n$ board so that there exists a unique partition of the board into $1 \times 2$ and $2 \times 1$ dominoes, none of which contains two marked cells.

## Geometry

G1. In a convex pentagon $A B C D E$, let $F$ be a point on $A C$ such that $\angle F B C=90^{\circ}$. Suppose triangles $A B F, A C D$ and $A D E$ are similar isosceles triangles with

$$
\angle F A B=\angle F B A=\angle D A C=\angle D C A=\angle E A D=\angle E D A .
$$

Let $M$ be the midpoint of $C F$. Point $X$ is chosen such that $A M X E$ is a parallelogram. Show that $B D, E M$ and $F X$ are concurrent.

G2. Let $A B C$ be a triangle with circumcircle $\Gamma$ and incentre $I$. Let $M$ be the midpoint of side $B C$. Denote by $D$ the foot of perpendicular from $I$ to side $B C$. The line through $I$ perpendicular to $A I$ meets sides $A B$ and $A C$ at $F$ and $E$ respectively. Suppose the circumcircle of triangle $A E F$ intersects $\Gamma$ at a point $X$ other than $A$. Prove that lines $X D$ and $A M$ meet on $\Gamma$.

G3. Let $B=(-1,0)$ and $C=(1,0)$ be fixed points on the coordinate plane. A nonempty, bounded subset $S$ of the plane is said to be nice if
(i) there is a point $T$ in $S$ such that for every point $Q$ in $S$, the segment $T Q$ lies entirely in $S$; and
(ii) for any triangle $P_{1} P_{2} P_{3}$, there exists a unique point $A$ in $S$ and a permutation $\sigma$ of the indices $\{1,2,3\}$ for which triangles $A B C$ and $P_{\sigma(1)} P_{\sigma(2)} P_{\sigma(3)}$ are similar.

Prove that there exist two distinct nice subsets $S$ and $S^{\prime}$ of the set $\{(x, y): x \geqslant 0, y \geqslant 0\}$ such that if $A \in S$ and $A^{\prime} \in S^{\prime}$ are the unique choices of points in (ii), then the product $B A \cdot B A^{\prime}$ is a constant independent of the triangle $P_{1} P_{2} P_{3}$.

G4. Let $A B C$ be a triangle with $A B=A C \neq B C$ and let $I$ be its incentre. The line $B I$ meets $A C$ at $D$, and the line through $D$ perpendicular to $A C$ meets $A I$ at $E$. Prove that the reflection of $I$ in $A C$ lies on the circumcircle of triangle $B D E$.

G5. Let $D$ be the foot of perpendicular from $A$ to the Euler line (the line passing through the circumcentre and the orthocentre) of an acute scalene triangle $A B C$. A circle $\omega$ with centre $S$ passes through $A$ and $D$, and it intersects sides $A B$ and $A C$ at $X$ and $Y$ respectively. Let $P$ be the foot of altitude from $A$ to $B C$, and let $M$ be the midpoint of $B C$. Prove that the circumcentre of triangle $X S Y$ is equidistant from $P$ and $M$.

G6. Let $A B C D$ be a convex quadrilateral with $\angle A B C=\angle A D C<90^{\circ}$. The internal angle bisectors of $\angle A B C$ and $\angle A D C$ meet $A C$ at $E$ and $F$ respectively, and meet each other at point $P$. Let $M$ be the midpoint of $A C$ and let $\omega$ be the circumcircle of triangle $B P D$. Segments $B M$ and $D M$ intersect $\omega$ again at $X$ and $Y$ respectively. Denote by $Q$ the intersection point of lines $X E$ and $Y F$. Prove that $P Q \perp A C$.

G7. Let $I$ be the incentre of a non-equilateral triangle $A B C, I_{A}$ be the $A$-excentre, $I_{A}^{\prime}$ be the reflection of $I_{A}$ in $B C$, and $l_{A}$ be the reflection of line $A I_{A}^{\prime}$ in $A I$. Define points $I_{B}, I_{B}^{\prime}$ and line $l_{B}$ analogously. Let $P$ be the intersection point of $l_{A}$ and $l_{B}$.
(a) Prove that $P$ lies on line $O I$ where $O$ is the circumcentre of triangle $A B C$.
(b) Let one of the tangents from $P$ to the incircle of triangle $A B C$ meet the circumcircle at points $X$ and $Y$. Show that $\angle X I Y=120^{\circ}$.

G8. Let $A_{1}, B_{1}$ and $C_{1}$ be points on sides $B C, C A$ and $A B$ of an acute triangle $A B C$ respectively, such that $A A_{1}, B B_{1}$ and $C C_{1}$ are the internal angle bisectors of triangle $A B C$. Let $I$ be the incentre of triangle $A B C$, and $H$ be the orthocentre of triangle $A_{1} B_{1} C_{1}$. Show that

$$
A H+B H+C H \geqslant A I+B I+C I
$$

## Number Theory

N1. For any positive integer $k$, denote the sum of digits of $k$ in its decimal representation by $S(k)$. Find all polynomials $P(x)$ with integer coefficients such that for any positive integer $n \geqslant 2016$, the integer $P(n)$ is positive and

$$
S(P(n))=P(S(n))
$$

N2. Let $\tau(n)$ be the number of positive divisors of $n$. Let $\tau_{1}(n)$ be the number of positive divisors of $n$ which have remainders 1 when divided by 3 . Find all possible integral values of the fraction $\frac{\tau(10 n)}{\tau_{1}(10 n)}$.

N3. Define $P(n)=n^{2}+n+1$. For any positive integers $a$ and $b$, the set

$$
\{P(a), P(a+1), P(a+2), \ldots, P(a+b)\}
$$

is said to be fragrant if none of its elements is relatively prime to the product of the other elements. Determine the smallest size of a fragrant set.

N4. Let $n, m, k$ and $l$ be positive integers with $n \neq 1$ such that $n^{k}+m n^{l}+1$ divides $n^{k+l}-1$. Prove that

- $m=1$ and $l=2 k$; or
- $l \mid k$ and $m=\frac{n^{k-l}-1}{n^{l}-1}$.

N5. Let $a$ be a positive integer which is not a square number. Denote by $A$ the set of all positive integers $k$ such that

$$
\begin{equation*}
k=\frac{x^{2}-a}{x^{2}-y^{2}} \tag{1}
\end{equation*}
$$

for some integers $x$ and $y$ with $x>\sqrt{a}$. Denote by $B$ the set of all positive integers $k$ such that (1) is satisfied for some integers $x$ and $y$ with $0 \leqslant x<\sqrt{a}$. Prove that $A=B$.

N6. Denote by $\mathbb{N}$ the set of all positive integers. Find all functions $f: \mathbb{N} \rightarrow \mathbb{N}$ such that for all positive integers $m$ and $n$, the integer $f(m)+f(n)-m n$ is nonzero and divides $m f(m)+n f(n)$.

N7. Let $n$ be an odd positive integer. In the Cartesian plane, a cyclic polygon $P$ with area $S$ is chosen. All its vertices have integral coordinates, and all squares of its side lengths are divisible by $n$. Prove that $2 S$ is an integer divisible by $n$.

N8. Find all polynomials $P(x)$ of odd degree $d$ and with integer coefficients satisfying the following property: for each positive integer $n$, there exist $n$ positive integers $x_{1}, x_{2}, \ldots, x_{n}$ such that $\frac{1}{2}<\frac{P\left(x_{i}\right)}{P\left(x_{i}\right)}<2$ and $\frac{P\left(x_{i}\right)}{P\left(x_{j}\right)}$ is the $d$-th power of a rational number for every pair of indices $i$ and $j$ with $1 \leqslant i, j \leqslant n$.

## Solutions

## Algebra

A1. Let $a, b$ and $c$ be positive real numbers such that $\min \{a b, b c, c a\} \geqslant 1$. Prove that

$$
\begin{equation*}
\sqrt[3]{\left(a^{2}+1\right)\left(b^{2}+1\right)\left(c^{2}+1\right)} \leqslant\left(\frac{a+b+c}{3}\right)^{2}+1 \tag{1}
\end{equation*}
$$

Solution 1. We first show the following.

- Claim. For any positive real numbers $x, y$ with $x y \geqslant 1$, we have

$$
\begin{equation*}
\left(x^{2}+1\right)\left(y^{2}+1\right) \leqslant\left(\left(\frac{x+y}{2}\right)^{2}+1\right)^{2} \tag{2}
\end{equation*}
$$

Proof. Note that $x y \geqslant 1$ implies $\left(\frac{x+y}{2}\right)^{2}-1 \geqslant x y-1 \geqslant 0$. We find that $\left(x^{2}+1\right)\left(y^{2}+1\right)=(x y-1)^{2}+(x+y)^{2} \leqslant\left(\left(\frac{x+y}{2}\right)^{2}-1\right)^{2}+(x+y)^{2}=\left(\left(\frac{x+y}{2}\right)^{2}+1\right)^{2}$.

Without loss of generality, assume $a \geqslant b \geqslant c$. This implies $a \geqslant 1$. Let $d=\frac{a+b+c}{3}$. Note that

$$
a d=\frac{a(a+b+c)}{3} \geqslant \frac{1+1+1}{3}=1 .
$$

Then we can apply (2) to the pair $(a, d)$ and the pair $(b, c)$. We get

$$
\begin{equation*}
\left(a^{2}+1\right)\left(d^{2}+1\right)\left(b^{2}+1\right)\left(c^{2}+1\right) \leqslant\left(\left(\frac{a+d}{2}\right)^{2}+1\right)^{2}\left(\left(\frac{b+c}{2}\right)^{2}+1\right)^{2} \tag{3}
\end{equation*}
$$

Next, from

$$
\frac{a+d}{2} \cdot \frac{b+c}{2} \geqslant \sqrt{a d} \cdot \sqrt{b c} \geqslant 1
$$

we can apply (2) again to the pair $\left(\frac{a+d}{2}, \frac{b+c}{2}\right)$. Together with (3), we have

$$
\left(a^{2}+1\right)\left(d^{2}+1\right)\left(b^{2}+1\right)\left(c^{2}+1\right) \leqslant\left(\left(\frac{a+b+c+d}{4}\right)^{2}+1\right)^{4}=\left(d^{2}+1\right)^{4}
$$

Therefore, $\left(a^{2}+1\right)\left(b^{2}+1\right)\left(c^{2}+1\right) \leqslant\left(d^{2}+1\right)^{3}$, and (1) follows by taking cube root of both sides.

Comment. After justifying the Claim, one may also obtain (1) by mixing variables. Indeed, the function involved is clearly continuous, and hence it suffices to check that the condition $x y \geqslant 1$ is preserved under each mixing step. This is true since whenever $a b, b c, c a \geqslant 1$, we have

$$
\frac{a+b}{2} \cdot \frac{a+b}{2} \geqslant a b \geqslant 1 \quad \text { and } \quad \frac{a+b}{2} \cdot c \geqslant \frac{1+1}{2}=1 .
$$

Solution 2. Let $f(x)=\ln \left(1+x^{2}\right)$. Then the inequality (1) to be shown is equivalent to

$$
\frac{f(a)+f(b)+f(c)}{3} \leqslant f\left(\frac{a+b+c}{3}\right),
$$

while (2) becomes

$$
\frac{f(x)+f(y)}{2} \leqslant f\left(\frac{x+y}{2}\right)
$$

for $x y \geqslant 1$.
Without loss of generality, assume $a \geqslant b \geqslant c$. From the Claim in Solution 1, we have

$$
\frac{f(a)+f(b)+f(c)}{3} \leqslant \frac{f(a)+2 f\left(\frac{b+c}{2}\right)}{3} .
$$

Note that $a \geqslant 1$ and $\frac{b+c}{2} \geqslant \sqrt{b c} \geqslant 1$. Since

$$
f^{\prime \prime}(x)=\frac{2\left(1-x^{2}\right)}{\left(1+x^{2}\right)^{2}}
$$

we know that $f$ is concave on $[1, \infty)$. Then we can apply Jensen's Theorem to get

$$
\frac{f(a)+2 f\left(\frac{b+c}{2}\right)}{3} \leqslant f\left(\frac{a+2 \cdot \frac{b+c}{2}}{3}\right)=f\left(\frac{a+b+c}{3}\right) .
$$

This completes the proof.

A2. Find the smallest real constant $C$ such that for any positive real numbers $a_{1}, a_{2}, a_{3}, a_{4}$ and $a_{5}$ (not necessarily distinct), one can always choose distinct subscripts $i, j, k$ and $l$ such that

$$
\begin{equation*}
\left|\frac{a_{i}}{a_{j}}-\frac{a_{k}}{a_{l}}\right| \leqslant C . \tag{1}
\end{equation*}
$$

Answer. The smallest $C$ is $\frac{1}{2}$.
Solution. We first show that $C \leqslant \frac{1}{2}$. For any positive real numbers $a_{1} \leqslant a_{2} \leqslant a_{3} \leqslant a_{4} \leqslant a_{5}$, consider the five fractions

$$
\begin{equation*}
\frac{a_{1}}{a_{2}}, \frac{a_{3}}{a_{4}}, \frac{a_{1}}{a_{5}}, \frac{a_{2}}{a_{3}}, \frac{a_{4}}{a_{5}} . \tag{2}
\end{equation*}
$$

Each of them lies in the interval $(0,1]$. Therefore, by the Pigeonhole Principle, at least three of them must lie in $\left(0, \frac{1}{2}\right]$ or lie in $\left(\frac{1}{2}, 1\right]$ simultaneously. In particular, there must be two consecutive terms in (2) which belong to an interval of length $\frac{1}{2}$ (here, we regard $\frac{a_{1}}{a_{2}}$ and $\frac{a_{4}}{a_{5}}$ as consecutive). In other words, the difference of these two fractions is less than $\frac{1}{2}$. As the indices involved in these two fractions are distinct, we can choose them to be $i, j, k, l$ and conclude that $C \leqslant \frac{1}{2}$.

Next, we show that $C=\frac{1}{2}$ is best possible. Consider the numbers $1,2,2,2, n$ where $n$ is a large real number. The fractions formed by two of these numbers in ascending order are $\frac{1}{n}, \frac{2}{n}, \frac{1}{2}, \frac{2}{2}, \frac{2}{1}, \frac{n}{2}, \frac{n}{1}$. Since the indices $i, j, k, l$ are distinct, $\frac{1}{n}$ and $\frac{2}{n}$ cannot be chosen simultaneously. Therefore the minimum value of the left-hand side of (1) is $\frac{1}{2}-\frac{2}{n}$. When $n$ tends to infinity, this value approaches $\frac{1}{2}$, and so $C$ cannot be less than $\frac{1}{2}$.

These conclude that $C=\frac{1}{2}$ is the smallest possible choice.
Comment. The conclusion still holds if $a_{1}, a_{2}, \ldots, a_{5}$ are pairwise distinct, since in the construction, we may replace the 2 's by real numbers sufficiently close to 2 .

There are two possible simplifications for this problem:
(i) the answer $C=\frac{1}{2}$ is given to the contestants; or
(ii) simply ask the contestants to prove the inequality (1) for $C=\frac{1}{2}$.

A3. Find all integers $n \geqslant 3$ with the following property: for all real numbers $a_{1}, a_{2}, \ldots, a_{n}$ and $b_{1}, b_{2}, \ldots, b_{n}$ satisfying $\left|a_{k}\right|+\left|b_{k}\right|=1$ for $1 \leqslant k \leqslant n$, there exist $x_{1}, x_{2}, \ldots, x_{n}$, each of which is either -1 or 1 , such that

$$
\begin{equation*}
\left|\sum_{k=1}^{n} x_{k} a_{k}\right|+\left|\sum_{k=1}^{n} x_{k} b_{k}\right| \leqslant 1 \tag{1}
\end{equation*}
$$

Answer. $n$ can be any odd integer greater than or equal to 3 .
Solution 1. For any even integer $n \geqslant 4$, we consider the case

$$
a_{1}=a_{2}=\cdots=a_{n-1}=b_{n}=0 \quad \text { and } \quad b_{1}=b_{2}=\cdots=b_{n-1}=a_{n}=1
$$

The condition $\left|a_{k}\right|+\left|b_{k}\right|=1$ is satisfied for each $1 \leqslant k \leqslant n$. No matter how we choose each $x_{k}$, both sums $\sum_{k=1}^{n} x_{k} a_{k}$ and $\sum_{k=1}^{n} x_{k} b_{k}$ are odd integers. This implies $\left|\sum_{k=1}^{n} x_{k} a_{k}\right| \geqslant 1$ and $\left|\sum_{k=1}^{n} x_{k} b_{k}\right| \geqslant 1$, which shows (1) cannot hold.

For any odd integer $n \geqslant 3$, we may assume without loss of generality $b_{k} \geqslant 0$ for $1 \leqslant k \leqslant n$ (this can be done by flipping the pair $\left(a_{k}, b_{k}\right)$ to $\left(-a_{k},-b_{k}\right)$ and $x_{k}$ to $-x_{k}$ if necessary) and $a_{1} \geqslant a_{2} \geqslant \cdots \geqslant a_{m} \geqslant 0>a_{m+1} \geqslant \cdots \geqslant a_{n}$. We claim that the choice $x_{k}=(-1)^{k+1}$ for $1 \leqslant k \leqslant n$ will work. Define

$$
s=\sum_{k=1}^{m} x_{k} a_{k} \quad \text { and } \quad t=-\sum_{k=m+1}^{n} x_{k} a_{k} .
$$

Note that

$$
s=\left(a_{1}-a_{2}\right)+\left(a_{3}-a_{4}\right)+\cdots \geqslant 0
$$

by the assumption $a_{1} \geqslant a_{2} \geqslant \cdots \geqslant a_{m}$ (when $m$ is odd, there is a single term $a_{m}$ at the end, which is also positive). Next, we have

$$
s=a_{1}-\left(a_{2}-a_{3}\right)-\left(a_{4}-a_{5}\right)-\cdots \leqslant a_{1} \leqslant 1
$$

Similarly,

$$
t=\left(-a_{n}+a_{n-1}\right)+\left(-a_{n-2}+a_{n-3}\right)+\cdots \geqslant 0
$$

and

$$
t=-a_{n}+\left(a_{n-1}-a_{n-2}\right)+\left(a_{n-3}-a_{n-4}\right)+\cdots \leqslant-a_{n} \leqslant 1 .
$$

From the condition, we have $a_{k}+b_{k}=1$ for $1 \leqslant k \leqslant m$ and $-a_{k}+b_{k}=1$ for $m+1 \leqslant k \leqslant n$. It follows that $\sum_{k=1}^{n} x_{k} a_{k}=s-t$ and $\sum_{k=1}^{n} x_{k} b_{k}=1-s-t$. Hence it remains to prove

$$
|s-t|+|1-s-t| \leqslant 1
$$

under the constraint $0 \leqslant s, t \leqslant 1$. By symmetry, we may assume $s \geqslant t$. If $1-s-t \geqslant 0$, then we have

$$
|s-t|+|1-s-t|=s-t+1-s-t=1-2 t \leqslant 1
$$

If $1-s-t \leqslant 0$, then we have

$$
|s-t|+|1-s-t|=s-t-1+s+t=2 s-1 \leqslant 1
$$

Hence, the inequality is true in both cases.
These show $n$ can be any odd integer greater than or equal to 3 .

Solution 2. The even case can be handled in the same way as Solution 1. For the odd case, we prove by induction on $n$.

Firstly, for $n=3$, we may assume without loss of generality $a_{1} \geqslant a_{2} \geqslant a_{3} \geqslant 0$ and $b_{1}=a_{1}-1$ (if $b_{1}=1-a_{1}$, we may replace each $b_{k}$ by $-b_{k}$ ).

- Case 1. $b_{2}=a_{2}-1$ and $b_{3}=a_{3}-1$, in which case we take $\left(x_{1}, x_{2}, x_{3}\right)=(1,-1,1)$.

Let $c=a_{1}-a_{2}+a_{3}$ so that $0 \leqslant c \leqslant 1$. Then $\left|b_{1}-b_{2}+b_{3}\right|=\left|a_{1}-a_{2}+a_{3}-1\right|=1-c$ and hence $|c|+\left|b_{1}-b_{2}+b_{3}\right|=1$.

- Case 2. $b_{2}=1-a_{2}$ and $b_{3}=1-a_{3}$, in which case we take $\left(x_{1}, x_{2}, x_{3}\right)=(1,-1,1)$.

Let $c=a_{1}-a_{2}+a_{3}$ so that $0 \leqslant c \leqslant 1$. Since $a_{3} \leqslant a_{2}$ and $a_{1} \leqslant 1$, we have

$$
c-1 \leqslant b_{1}-b_{2}+b_{3}=a_{1}+a_{2}-a_{3}-1 \leqslant 1-c .
$$

This gives $\left|b_{1}-b_{2}+b_{3}\right| \leqslant 1-c$ and hence $|c|+\left|b_{1}-b_{2}+b_{3}\right| \leqslant 1$.

- Case 3. $b_{2}=a_{2}-1$ and $b_{3}=1-a_{3}$, in which case we take $\left(x_{1}, x_{2}, x_{3}\right)=(-1,1,1)$.

Let $c=-a_{1}+a_{2}+a_{3}$. If $c \geqslant 0$, then $a_{3} \leqslant 1$ and $a_{2} \leqslant a_{1}$ imply

$$
c-1 \leqslant-b_{1}+b_{2}+b_{3}=-a_{1}+a_{2}-a_{3}+1 \leqslant 1-c .
$$

If $c<0$, then $a_{1} \leqslant a_{2}+1$ and $a_{3} \geqslant 0$ imply

$$
-c-1 \leqslant-b_{1}+b_{2}+b_{3}=-a_{1}+a_{2}-a_{3}+1 \leqslant 1+c .
$$

In both cases, we get $\left|-b_{1}+b_{2}+b_{3}\right| \leqslant 1-|c|$ and hence $|c|+\left|-b_{1}+b_{2}+b_{3}\right| \leqslant 1$.

- Case 4. $b_{2}=1-a_{2}$ and $b_{3}=a_{3}-1$, in which case we take $\left(x_{1}, x_{2}, x_{3}\right)=(-1,1,1)$.

Let $c=-a_{1}+a_{2}+a_{3}$. If $c \geqslant 0$, then $a_{2} \leqslant 1$ and $a_{3} \leqslant a_{1}$ imply

$$
c-1 \leqslant-b_{1}+b_{2}+b_{3}=-a_{1}-a_{2}+a_{3}+1 \leqslant 1-c .
$$

If $c<0$, then $a_{1} \leqslant a_{3}+1$ and $a_{2} \geqslant 0$ imply

$$
-c-1 \leqslant-b_{1}+b_{2}+b_{3}=-a_{1}-a_{2}+a_{3}+1 \leqslant 1+c .
$$

In both cases, we get $\left|-b_{1}+b_{2}+b_{3}\right| \leqslant 1-|c|$ and hence $|c|+\left|-b_{1}+b_{2}+b_{3}\right| \leqslant 1$.
We have found $x_{1}, x_{2}, x_{3}$ satisfying (1) in each case for $n=3$.
Now, let $n \geqslant 5$ be odd and suppose the result holds for any smaller odd cases. Again we may assume $a_{k} \geqslant 0$ for each $1 \leqslant k \leqslant n$. By the Pigeonhole Principle, there are at least three indices $k$ for which $b_{k}=a_{k}-1$ or $b_{k}=1-a_{k}$. Without loss of generality, suppose $b_{k}=a_{k}-1$ for $k=1,2,3$. Again by the Pigeonhole Principle, as $a_{1}, a_{2}, a_{3}$ lies between 0 and 1 , the difference of two of them is at most $\frac{1}{2}$. By changing indices if necessary, we may assume $0 \leqslant d=a_{1}-a_{2} \leqslant \frac{1}{2}$.

By the inductive hypothesis, we can choose $x_{3}, x_{4}, \ldots, x_{n}$ such that $a^{\prime}=\sum_{k=3}^{n} x_{k} a_{k}$ and $b^{\prime}=\sum_{k=3}^{n} x_{k} b_{k}$ satisfy $\left|a^{\prime}\right|+\left|b^{\prime}\right| \leqslant 1$. We may further assume $a^{\prime} \geqslant 0$.

- Case 1. $b^{\prime} \geqslant 0$, in which case we take $\left(x_{1}, x_{2}\right)=(-1,1)$.

We have $\left|-a_{1}+a_{2}+a^{\prime}\right|+\left|-\left(a_{1}-1\right)+\left(a_{2}-1\right)+b^{\prime}\right|=\left|-d+a^{\prime}\right|+\left|-d+b^{\prime}\right| \leqslant$ $\max \left\{a^{\prime}+b^{\prime}-2 d, a^{\prime}-b^{\prime}, b^{\prime}-a^{\prime}, 2 d-a^{\prime}-b^{\prime}\right\} \leqslant 1$ since $0 \leqslant a^{\prime}, b^{\prime}, a^{\prime}+b^{\prime} \leqslant 1$ and $0 \leqslant d \leqslant \frac{1}{2}$.

- Case 2. $0>b^{\prime} \geqslant-a^{\prime}$, in which case we take $\left(x_{1}, x_{2}\right)=(-1,1)$.

We have $\left|-a_{1}+a_{2}+a^{\prime}\right|+\left|-\left(a_{1}-1\right)+\left(a_{2}-1\right)+b^{\prime}\right|=\left|-d+a^{\prime}\right|+\left|-d+b^{\prime}\right|$. If $-d+a^{\prime} \geqslant 0$, this equals $a^{\prime}-b^{\prime}=\left|a^{\prime}\right|+\left|b^{\prime}\right| \leqslant 1$. If $-d+a^{\prime}<0$, this equals $2 d-a^{\prime}-b^{\prime} \leqslant 2 d \leqslant 1$.

- Case 3. $b^{\prime}<-a^{\prime}$, in which case we take $\left(x_{1}, x_{2}\right)=(1,-1)$.

We have $\left|a_{1}-a_{2}+a^{\prime}\right|+\left|\left(a_{1}-1\right)-\left(a_{2}-1\right)+b^{\prime}\right|=\left|d+a^{\prime}\right|+\left|d+b^{\prime}\right|$. If $d+b^{\prime} \geqslant 0$, this equals $2 d+a^{\prime}+b^{\prime}<2 d \leqslant 1$. If $d+b^{\prime}<0$, this equals $a^{\prime}-b^{\prime}=\left|a^{\prime}\right|+\left|b^{\prime}\right| \leqslant 1$.

Therefore, we have found $x_{1}, x_{2}, \ldots, x_{n}$ satisfying (1) in each case. By induction, the property holds for all odd integers $n \geqslant 3$.

A4. Denote by $\mathbb{R}^{+}$the set of all positive real numbers. Find all functions $f: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$such that

$$
\begin{equation*}
x f\left(x^{2}\right) f(f(y))+f(y f(x))=f(x y)\left(f\left(f\left(x^{2}\right)\right)+f\left(f\left(y^{2}\right)\right)\right) \tag{1}
\end{equation*}
$$

for all positive real numbers $x$ and $y$.
Answer. $f(x)=\frac{1}{x}$ for any $x \in \mathbb{R}^{+}$.
Solution 1. Taking $x=y=1$ in (1), we get $f(1) f(f(1))+f(f(1))=2 f(1) f(f(1))$ and hence $f(1)=1$. Swapping $x$ and $y$ in (1) and comparing with (1) again, we find

$$
\begin{equation*}
x f\left(x^{2}\right) f(f(y))+f(y f(x))=y f\left(y^{2}\right) f(f(x))+f(x f(y)) . \tag{2}
\end{equation*}
$$

Taking $y=1$ in (2), we have $x f\left(x^{2}\right)+f(f(x))=f(f(x))+f(x)$, that is,

$$
\begin{equation*}
f\left(x^{2}\right)=\frac{f(x)}{x} \tag{3}
\end{equation*}
$$

Take $y=1$ in (1) and apply (3) to $x f\left(x^{2}\right)$. We get $f(x)+f(f(x))=f(x)\left(f\left(f\left(x^{2}\right)\right)+1\right)$, which implies

$$
\begin{equation*}
f\left(f\left(x^{2}\right)\right)=\frac{f(f(x))}{f(x)} \tag{4}
\end{equation*}
$$

For any $x \in \mathbb{R}^{+}$, we find that

$$
\begin{equation*}
f\left(f(x)^{2}\right) \stackrel{(3)}{=} \frac{f(f(x))}{f(x)} \stackrel{(4)}{=} f\left(f\left(x^{2}\right)\right) \stackrel{(3)}{=} f\left(\frac{f(x)}{x}\right) \tag{5}
\end{equation*}
$$

It remains to show the following key step.

- Claim. The function $f$ is injective.

Proof. Using (3) and (4), we rewrite (1) as

$$
\begin{equation*}
f(x) f(f(y))+f(y f(x))=f(x y)\left(\frac{f(f(x))}{f(x)}+\frac{f(f(y))}{f(y)}\right) . \tag{6}
\end{equation*}
$$

Take $x=y$ in (6) and apply (3). This gives $f(x) f(f(x))+f(x f(x))=2 \frac{f(f(x))}{x}$, which means

$$
\begin{equation*}
f(x f(x))=f(f(x))\left(\frac{2}{x}-f(x)\right) . \tag{7}
\end{equation*}
$$

Using (3), equation (2) can be rewritten as

$$
\begin{equation*}
f(x) f(f(y))+f(y f(x))=f(y) f(f(x))+f(x f(y)) \tag{8}
\end{equation*}
$$

Suppose $f(x)=f(y)$ for some $x, y \in \mathbb{R}^{+}$. Then (8) implies

$$
f(y f(y))=f(y f(x))=f(x f(y))=f(x f(x)) .
$$

Using (7), this gives

$$
f(f(y))\left(\frac{2}{y}-f(y)\right)=f(f(x))\left(\frac{2}{x}-f(x)\right) .
$$

Noting $f(x)=f(y)$, we find $x=y$. This establishes the injectivity.

By the Claim and (5), we get the only possible solution $f(x)=\frac{1}{x}$. It suffices to check that this is a solution. Indeed, the left-hand side of (1) becomes

$$
x \cdot \frac{1}{x^{2}} \cdot y+\frac{x}{y}=\frac{y}{x}+\frac{x}{y},
$$

while the right-hand side becomes

$$
\frac{1}{x y}\left(x^{2}+y^{2}\right)=\frac{x}{y}+\frac{y}{x} .
$$

The two sides agree with each other.
Solution 2. Taking $x=y=1$ in (1), we get $f(1) f(f(1))+f(f(1))=2 f(1) f(f(1))$ and hence $f(1)=1$. Putting $x=1$ in (1), we have $f(f(y))+f(y)=f(y)\left(1+f\left(f\left(y^{2}\right)\right)\right)$ so that

$$
\begin{equation*}
f(f(y))=f(y) f\left(f\left(y^{2}\right)\right) \tag{9}
\end{equation*}
$$

Putting $y=1$ in (1), we get $x f\left(x^{2}\right)+f(f(x))=f(x)\left(f\left(f\left(x^{2}\right)\right)+1\right)$. Using (9), this gives

$$
\begin{equation*}
x f\left(x^{2}\right)=f(x) \tag{10}
\end{equation*}
$$

Replace $y$ by $\frac{1}{x}$ in (1). Then we have

$$
x f\left(x^{2}\right) f\left(f\left(\frac{1}{x}\right)\right)+f\left(\frac{f(x)}{x}\right)=f\left(f\left(x^{2}\right)\right)+f\left(f\left(\frac{1}{x^{2}}\right)\right) .
$$

The relation (10) shows $f\left(\frac{f(x)}{x}\right)=f\left(f\left(x^{2}\right)\right)$. Also, using (9) with $y=\frac{1}{x}$ and using (10) again, the last equation reduces to

$$
\begin{equation*}
f(x) f\left(\frac{1}{x}\right)=1 \tag{11}
\end{equation*}
$$

Replace $x$ by $\frac{1}{x}$ and $y$ by $\frac{1}{y}$ in (1) and apply (11). We get

$$
\frac{1}{x f\left(x^{2}\right) f(f(y))}+\frac{1}{f(y f(x))}=\frac{1}{f(x y)}\left(\frac{1}{f\left(f\left(x^{2}\right)\right)}+\frac{1}{f\left(f\left(y^{2}\right)\right)}\right) .
$$

Clearing denominators, we can use (1) to simplify the numerators and obtain

$$
f(x y)^{2} f\left(f\left(x^{2}\right)\right) f\left(f\left(y^{2}\right)\right)=x f\left(x^{2}\right) f(f(y)) f(y f(x)) .
$$

Using (9) and (10), this is the same as

$$
\begin{equation*}
f(x y)^{2} f(f(x))=f(x)^{2} f(y) f(y f(x)) \tag{12}
\end{equation*}
$$

Substitute $y=f(x)$ in (12) and apply (10) (with $x$ replaced by $f(x)$ ). We have

$$
\begin{equation*}
f(x f(x))^{2}=f(x) f(f(x)) \tag{13}
\end{equation*}
$$

Taking $y=x$ in (12), squaring both sides, and using (10) and (13), we find that

$$
\begin{equation*}
f(f(x))=x^{4} f(x)^{3} . \tag{14}
\end{equation*}
$$

Finally, we combine (9), (10) and (14) to get

$$
y^{4} f(y)^{3} \stackrel{(14)}{=} f(f(y)) \stackrel{(9)}{=} f(y) f\left(f\left(y^{2}\right)\right) \stackrel{(14)}{=} f(y) y^{8} f\left(y^{2}\right)^{3} \stackrel{(10)}{=} y^{5} f(y)^{4}
$$

which implies $f(y)=\frac{1}{y}$. This is a solution by the checking in Solution 1.

## A5.

(a) Prove that for every positive integer $n$, there exists a fraction $\frac{a}{b}$ where $a$ and $b$ are integers satisfying $0<b \leqslant \sqrt{n}+1$ and $\sqrt{n} \leqslant \frac{a}{b} \leqslant \sqrt{n+1}$.
(b) Prove that there are infinitely many positive integers $n$ such that there is no fraction $\frac{a}{b}$ where $a$ and $b$ are integers satisfying $0<b \leqslant \sqrt{n}$ and $\sqrt{n} \leqslant \frac{a}{b} \leqslant \sqrt{n+1}$.

## Solution.

(a) Let $r$ be the unique positive integer for which $r^{2} \leqslant n<(r+1)^{2}$. Write $n=r^{2}+s$. Then we have $0 \leqslant s \leqslant 2 r$. We discuss in two cases according to the parity of $s$.

- Case 1. $s$ is even.

Consider the number $\left(r+\frac{s}{2 r}\right)^{2}=r^{2}+s+\left(\frac{s}{2 r}\right)^{2}$. We find that

$$
n=r^{2}+s \leqslant r^{2}+s+\left(\frac{s}{2 r}\right)^{2} \leqslant r^{2}+s+1=n+1
$$

It follows that

$$
\sqrt{n} \leqslant r+\frac{s}{2 r} \leqslant \sqrt{n+1}
$$

Since $s$ is even, we can choose the fraction $r+\frac{s}{2 r}=\frac{r^{2}+(s / 2)}{r}$ since $r \leqslant \sqrt{n}$.

- Case 2. $s$ is odd.

Consider the number $\left(r+1-\frac{2 r+1-s}{2(r+1)}\right)^{2}=(r+1)^{2}-(2 r+1-s)+\left(\frac{2 r+1-s}{2(r+1)}\right)^{2}$. We find that

$$
\begin{aligned}
n=r^{2}+s=(r+1)^{2}-(2 r+1-s) & \leqslant(r+1)^{2}-(2 r+1-s)+\left(\frac{2 r+1-s}{2(r+1)}\right)^{2} \\
& \leqslant(r+1)^{2}-(2 r+1-s)+1=n+1
\end{aligned}
$$

It follows that

$$
\sqrt{n} \leqslant r+1-\frac{2 r+1-s}{2(r+1)} \leqslant \sqrt{n+1}
$$

Since $s$ is odd, we can choose the fraction $(r+1)-\frac{2 r+1-s}{2(r+1)}=\frac{(r+1)^{2}-r+((s-1) / 2)}{r+1}$ since $r+1 \leqslant \sqrt{n}+1$.
(b) We show that for every positive integer $r$, there is no fraction $\frac{a}{b}$ with $b \leqslant \sqrt{r^{2}+1}$ such that $\sqrt{r^{2}+1} \leqslant \frac{a}{b} \leqslant \sqrt{r^{2}+2}$. Suppose on the contrary that such a fraction exists. Since $b \leqslant \sqrt{r^{2}+1}<r+1$ and $b$ is an integer, we have $b \leqslant r$. Hence,

$$
(b r)^{2}<b^{2}\left(r^{2}+1\right) \leqslant a^{2} \leqslant b^{2}\left(r^{2}+2\right) \leqslant b^{2} r^{2}+2 b r<(b r+1)^{2}
$$

This shows the square number $a^{2}$ is strictly bounded between the two consecutive squares $(b r)^{2}$ and $(b r+1)^{2}$, which is impossible. Hence, we have found infinitely many $n=r^{2}+1$ for which there is no fraction of the desired form.

A6. The equation

$$
(x-1)(x-2) \cdots(x-2016)=(x-1)(x-2) \cdots(x-2016)
$$

is written on the board. One tries to erase some linear factors from both sides so that each side still has at least one factor, and the resulting equation has no real roots. Find the least number of linear factors one needs to erase to achieve this.

Answer. 2016.
Solution. Since there are 2016 common linear factors on both sides, we need to erase at least 2016 factors. We claim that the equation has no real roots if we erase all factors $(x-k)$ on the left-hand side with $k \equiv 2,3(\bmod 4)$, and all factors $(x-m)$ on the right-hand side with $m \equiv 0,1(\bmod 4)$. Therefore, it suffices to show that no real number $x$ satisfies

$$
\begin{equation*}
\prod_{j=0}^{503}(x-4 j-1)(x-4 j-4)=\prod_{j=0}^{503}(x-4 j-2)(x-4 j-3) \tag{1}
\end{equation*}
$$

- Case 1. $x=1,2, \ldots, 2016$.

In this case, one side of (1) is zero while the other side is not. This shows $x$ cannot satisfy (1).

- Case 2. $4 k+1<x<4 k+2$ or $4 k+3<x<4 k+4$ for some $k=0,1, \ldots, 503$.

For $j=0,1, \ldots, 503$ with $j \neq k$, the product $(x-4 j-1)(x-4 j-4)$ is positive. For $j=k$, the product $(x-4 k-1)(x-4 k-4)$ is negative. This shows the left-hand side of (1) is negative. On the other hand, each product $(x-4 j-2)(x-4 j-3)$ on the right-hand side of (1) is positive. This yields a contradiction.

- Case 3. $x<1$ or $x>2016$ or $4 k<x<4 k+1$ for some $k=1,2, \ldots, 503$.

The equation (1) can be rewritten as

$$
1=\prod_{j=0}^{503} \frac{(x-4 j-1)(x-4 j-4)}{(x-4 j-2)(x-4 j-3)}=\prod_{j=0}^{503}\left(1-\frac{2}{(x-4 j-2)(x-4 j-3)}\right)
$$

Note that $(x-4 j-2)(x-4 j-3)>2$ for $0 \leqslant j \leqslant 503$ in this case. So each term in the product lies strictly between 0 and 1 , and the whole product must be less than 1 , which is impossible.

- Case 4. $4 k+2<x<4 k+3$ for some $k=0,1, \ldots, 503$.

This time we rewrite (1) as

$$
\begin{aligned}
1 & =\frac{x-1}{x-2} \cdot \frac{x-2016}{x-2015} \prod_{j=1}^{503} \frac{(x-4 j)(x-4 j-1)}{(x-4 j+1)(x-4 j-2)} \\
& =\frac{x-1}{x-2} \cdot \frac{x-2016}{x-2015} \prod_{j=1}^{503}\left(1+\frac{2}{(x-4 j+1)(x-4 j-2)}\right)
\end{aligned}
$$

Clearly, $\frac{x-1}{x-2}$ and $\frac{x-2016}{x-2015}$ are both greater than 1. For the range of $x$ in this case, each term in the product is also greater than 1 . Then the right-hand side must be greater than 1 and hence a contradiction arises.

From the four cases, we conclude that (1) has no real roots. Hence, the minimum number of linear factors to be erased is 2016 .

Comment. We discuss the general case when 2016 is replaced by a positive integer $n$. The above solution works equally well when $n$ is divisible by 4 .

If $n \equiv 2(\bmod 4)$, one may leave $l(x)=(x-1)(x-2) \cdots\left(x-\frac{n}{2}\right)$ on the left-hand side and $r(x)=\left(x-\frac{n}{2}-1\right)\left(x-\frac{n}{2}-2\right) \cdots(x-n)$ on the right-hand side. One checks that for $x<\frac{n+1}{2}$, we have $|l(x)|<|r(x)|$, while for $x>\frac{n+1}{2}$, we have $|l(x)|>|r(x)|$.

If $n \equiv 3(\bmod 4)$, one may leave $l(x)=(x-1)(x-2) \cdots\left(x-\frac{n+1}{2}\right)$ on the left-hand side and $r(x)=\left(x-\frac{n+3}{2}\right)\left(x-\frac{x+5}{2}\right) \cdots(x-n)$ on the right-hand side. For $x<1$ or $\frac{n+1}{2}<x<\frac{n+3}{2}$, we have $l(x)>0>r(x)$. For $1<x<\frac{n+1}{2}$, we have $|l(x)|<|r(x)|$. For $x>\frac{n+3}{2}$, we have $|l(x)|>|r(x)|$.

If $n \equiv 1(\bmod 4)$, as the proposer mentioned, the situation is a bit more out of control. Since the construction for $n-1 \equiv 0(\bmod 4)$ works, the answer can be either $n$ or $n-1$. For $n=5$, we can leave the products $(x-1)(x-2)(x-3)(x-4)$ and $(x-5)$. For $n=9$, the only example that works is $l(x)=(x-1)(x-2)(x-9)$ and $r(x)=(x-3)(x-4) \cdots(x-8)$, while there seems to be no such partition for $n=13$.

A7. Denote by $\mathbb{R}$ the set of all real numbers. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ such that $f(0) \neq 0$ and

$$
\begin{equation*}
f(x+y)^{2}=2 f(x) f(y)+\max \left\{f\left(x^{2}\right)+f\left(y^{2}\right), f\left(x^{2}+y^{2}\right)\right\} \tag{1}
\end{equation*}
$$

for all real numbers $x$ and $y$.

## Answer.

- $f(x)=-1$ for any $x \in \mathbb{R}$; or
- $f(x)=x-1$ for any $x \in \mathbb{R}$.

Solution 1. Taking $x=y=0$ in (1), we get $f(0)^{2}=2 f(0)^{2}+\max \{2 f(0), f(0)\}$. If $f(0)>0$, then $f(0)^{2}+2 f(0)=0$ gives no positive solution. If $f(0)<0$, then $f(0)^{2}+f(0)=0$ gives $f(0)=-1$. Putting $y=0$ in (1), we have $f(x)^{2}=-2 f(x)+f\left(x^{2}\right)$, which is the same as $(f(x)+1)^{2}=f\left(x^{2}\right)+1$. Let $g(x)=f(x)+1$. Then for any $x \in \mathbb{R}$, we have

$$
\begin{equation*}
g\left(x^{2}\right)=g(x)^{2} \geqslant 0 \tag{2}
\end{equation*}
$$

From (1), we find that $f(x+y)^{2} \geqslant 2 f(x) f(y)+f\left(x^{2}\right)+f\left(y^{2}\right)$. In terms of $g$, this becomes $(g(x+y)-1)^{2} \geqslant 2(g(x)-1)(g(y)-1)+g\left(x^{2}\right)+g\left(y^{2}\right)-2$. Using (2), this means

$$
\begin{equation*}
(g(x+y)-1)^{2} \geqslant(g(x)+g(y)-1)^{2}-1 . \tag{3}
\end{equation*}
$$

Putting $x=1$ in (2), we get $g(1)=0$ or 1 . The two cases are handled separately.

- Case 1. $g(1)=0$, which is the same as $f(1)=-1$.

We put $x=-1$ and $y=0$ in (1). This gives $f(-1)^{2}=-2 f(-1)-1$, which forces $f(-1)=-1$. Next, we take $x=-1$ and $y=1$ in (1) to get $1=2+\max \{-2, f(2)\}$. This clearly implies $1=2+f(2)$ and hence $f(2)=-1$, that is, $g(2)=0$. From (2), we can prove inductively that $g\left(2^{2^{n}}\right)=g(2)^{2^{n}}=0$ for any $n \in \mathbb{N}$. Substitute $y=2^{2^{n}}-x$ in (3). We obtain

$$
\left(g(x)+g\left(2^{2^{n}}-x\right)-1\right)^{2} \leqslant\left(g\left(2^{2^{n}}\right)-1\right)^{2}+1=2
$$

For any fixed $x \geqslant 0$, we consider $n$ to be sufficiently large so that $2^{2^{n}}-x>0$. From (2), this implies $g\left(2^{2^{n}}-x\right) \geqslant 0$ so that $g(x) \leqslant 1+\sqrt{2}$. Using (2) again, we get

$$
g(x)^{2^{n}}=g\left(x^{2^{n}}\right) \leqslant 1+\sqrt{2}
$$

for any $n \in \mathbb{N}$. Therefore, $|g(x)| \leqslant 1$ for any $x \geqslant 0$.
If there exists $a \in \mathbb{R}$ for which $g(a) \neq 0$, then for sufficiently large $n$ we must have $g\left(\left(a^{2}\right)^{\frac{1}{2^{n}}}\right)=g\left(a^{2}\right)^{\frac{1}{2^{n}}}>\frac{1}{2}$. By taking $x=-y=-\left(a^{2}\right)^{\frac{1}{2^{n}}}$ in (1), we obtain

$$
\begin{aligned}
1 & =2 f(x) f(-x)+\max \left\{2 f\left(x^{2}\right), f\left(2 x^{2}\right)\right\} \\
& =2(g(x)-1)(g(-x)-1)+\max \left\{2\left(g\left(x^{2}\right)-1\right), g\left(2 x^{2}\right)-1\right\} \\
& \leqslant 2\left(-\frac{1}{2}\right)\left(-\frac{1}{2}\right)+0=\frac{1}{2}
\end{aligned}
$$

since $|g(-x)|=|g(x)| \in\left(\frac{1}{2}, 1\right]$ by (2) and the choice of $x$, and since $g(z) \leqslant 1$ for $z \geqslant 0$. This yields a contradiction and hence $g(x)=0$ must hold for any $x$. This means $f(x)=-1$ for any $x \in \mathbb{R}$, which clearly satisfies (1).

- Case 2. $g(1)=1$, which is the same as $f(1)=0$.

We put $x=-1$ and $y=1$ in (1) to get $1=\max \{0, f(2)\}$. This clearly implies $f(2)=1$ and hence $g(2)=2$. Setting $x=2 n$ and $y=2$ in (3), we have

$$
(g(2 n+2)-1)^{2} \geqslant(g(2 n)+1)^{2}-1
$$

By induction on $n$, it is easy to prove that $g(2 n) \geqslant n+1$ for all $n \in \mathbb{N}$. For any real number $a>1$, we choose a large $n \in \mathbb{N}$ and take $k$ to be the positive integer such that $2 k \leqslant a^{2^{n}}<2 k+2$. From (2) and (3), we have

$$
\left(g(a)^{2^{n}}-1\right)^{2}+1=\left(g\left(a^{2^{n}}\right)-1\right)^{2}+1 \geqslant\left(g(2 k)+g\left(a^{2^{n}}-2 k\right)-1\right)^{2} \geqslant k^{2}>\frac{1}{4}\left(a^{2^{n}}-2\right)^{2}
$$

since $g\left(a^{2^{n}}-2 k\right) \geqslant 0$. For large $n$, this clearly implies $g(a)^{2^{n}}>1$. Thus,

$$
\left(g(a)^{2^{n}}\right)^{2}>\left(g(a)^{2^{n}}-1\right)^{2}+1>\frac{1}{4}\left(a^{2^{n}}-2\right)^{2}
$$

This yields

$$
\begin{equation*}
g(a)^{2^{n}}>\frac{1}{2}\left(a^{2^{n}}-2\right) . \tag{4}
\end{equation*}
$$

Note that

$$
\frac{a^{2^{n}}}{a^{2^{n}}-2}=1+\frac{2}{a^{2^{n}}-2} \leqslant\left(1+\frac{2}{2^{n}\left(a^{2^{n}}-2\right)}\right)^{2^{n}}
$$

by binomial expansion. This can be rewritten as

$$
\left(a^{2^{n}}-2\right)^{\frac{1}{2^{n}}} \geqslant \frac{a}{1+\frac{2}{2^{n}\left(a^{2^{n}}-2\right)}}
$$

Together with (4), we conclude $g(a) \geqslant a$ by taking $n$ sufficiently large.
Consider $x=n a$ and $y=a>1$ in (3). This gives $(g((n+1) a)-1)^{2} \geqslant(g(n a)+g(a)-1)^{2}-1$. By induction on $n$, it is easy to show $g(n a) \geqslant(n-1)(g(a)-1)+a$ for any $n \in \mathbb{N}$. We choose a large $n \in \mathbb{N}$ and take $k$ to be the positive integer such that $k a \leqslant 2^{2^{n}}<(k+1) a$. Using (2) and (3), we have
$2^{2^{n+1}}>\left(2^{2^{n}}-1\right)^{2}+1=\left(g\left(2^{2^{n}}\right)-1\right)^{2}+1 \geqslant\left(g\left(2^{2^{n}}-k a\right)+g(k a)-1\right)^{2} \geqslant((k-1)(g(a)-1)+a-1)^{2}$, from which it follows that

$$
2^{2^{n}} \geqslant(k-1)(g(a)-1)+a-1>\frac{2^{2^{n}}}{a}(g(a)-1)-2(g(a)-1)+a-1
$$

holds for sufficiently large $n$. Hence, we must have $\frac{g(a)-1}{a} \leqslant 1$, which implies $g(a) \leqslant a+1$ for any $a>1$. Then for large $n \in \mathbb{N}$, from (3) and (2) we have

$$
4 a^{2^{n+1}}=\left(2 a^{2^{n}}\right)^{2} \geqslant\left(g\left(2 a^{2^{n}}\right)-1\right)^{2} \geqslant\left(2 g\left(a^{2^{n}}\right)-1\right)^{2}-1=\left(2 g(a)^{2^{n}}-1\right)^{2}-1
$$

This implies

$$
2 a^{2^{n}}>\frac{1}{2}\left(1+\sqrt{4 a^{2 n+1}+1}\right) \geqslant g(a)^{2^{n}}
$$

When $n$ tends to infinity, this forces $g(a) \leqslant a$. Together with $g(a) \geqslant a$, we get $g(a)=a$ for all real numbers $a>1$, that is, $f(a)=a-1$ for all $a>1$.

Finally, for any $x \in \mathbb{R}$, we choose $y$ sufficiently large in (1) so that $y, x+y>1$. This gives $(x+y-1)^{2}=2 f(x)(y-1)+\max \left\{f\left(x^{2}\right)+y^{2}-1, x^{2}+y^{2}-1\right\}$, which can be rewritten as

$$
2(x-1-f(x)) y=-x^{2}+2 x-2-2 f(x)+\max \left\{f\left(x^{2}\right), x^{2}\right\} .
$$

As the right-hand side is fixed, this can only hold for all large $y$ when $f(x)=x-1$. We now check that this function satisfies (1). Indeed, we have

$$
\begin{aligned}
f(x+y)^{2} & =(x+y-1)^{2}=2(x-1)(y-1)+\left(x^{2}+y^{2}-1\right) \\
& =2 f(x) f(y)+\max \left\{f\left(x^{2}\right)+f\left(y^{2}\right), f\left(x^{2}+y^{2}\right)\right\} .
\end{aligned}
$$

Solution 2. Taking $x=y=0$ in (1), we get $f(0)^{2}=2 f(0)^{2}+\max \{2 f(0), f(0)\}$. If $f(0)>0$, then $f(0)^{2}+2 f(0)=0$ gives no positive solution. If $f(0)<0$, then $f(0)^{2}+f(0)=0$ gives $f(0)=-1$. Putting $y=0$ in (1), we have

$$
\begin{equation*}
f(x)^{2}=-2 f(x)+f\left(x^{2}\right) . \tag{5}
\end{equation*}
$$

Replace $x$ by $-x$ in (5) and compare with (5) again. We get $f(x)^{2}+2 f(x)=f(-x)^{2}+2 f(-x)$, which implies

$$
\begin{equation*}
f(x)=f(-x) \quad \text { or } \quad f(x)+f(-x)=-2 \tag{6}
\end{equation*}
$$

Taking $x=y$ and $x=-y$ respectively in (1) and comparing the two equations obtained, we have

$$
\begin{equation*}
f(2 x)^{2}-2 f(x)^{2}=1-2 f(x) f(-x) \tag{7}
\end{equation*}
$$

Combining (6) and (7) to eliminate $f(-x)$, we find that $f(2 x)$ can be $\pm 1$ (when $f(x)=f(-x))$ or $\pm(2 f(x)+1)$ (when $f(x)+f(-x)=-2)$.

We prove the following.

- Claim. $f(x)+f(-x)=-2$ for any $x \in \mathbb{R}$.

Proof. Suppose there exists $a \in \mathbb{R}$ such that $f(a)+f(-a) \neq-2$. Then $f(a)=f(-a) \neq-1$ and we may assume $a>0$. We first show that $f(a) \neq 1$. Suppose $f(a)=1$. Consider $y=a$ in (7). We get $f(2 a)^{2}=1$. Taking $x=y=a$ in (1), we have $1=2+\max \left\{2 f\left(a^{2}\right), f\left(2 a^{2}\right)\right\}$. From (5), $f\left(a^{2}\right)=3$ so that $1 \geqslant 2+6$. This is impossible, and thus $f(a) \neq 1$.

As $f(a) \neq \pm 1$, we have $f(a)= \pm\left(2 f\left(\frac{a}{2}\right)+1\right)$. Similarly, $f(-a)= \pm\left(2 f\left(-\frac{a}{2}\right)+1\right)$. These two expressions are equal since $f(a)=f(-a)$. If $f\left(\frac{a}{2}\right)=f\left(-\frac{a}{2}\right)$, then the above argument works when we replace $a$ by $\frac{a}{2}$. In particular, we have $f(a)^{2}=f\left(2 \cdot \frac{a}{2}\right)^{2}=1$, which is a contradiction. Therefore, (6) forces $f\left(\frac{a}{2}\right)+f\left(-\frac{a}{2}\right)=-2$. Then we get

$$
\pm\left(2 f\left(\frac{a}{2}\right)+1\right)= \pm\left(-2 f\left(\frac{a}{2}\right)-3\right) .
$$

For any choices of the two signs, we either get a contradiction or $f\left(\frac{a}{2}\right)=-1$, in which case $f\left(\frac{a}{2}\right)=f\left(-\frac{a}{2}\right)$ and hence $f(a)= \pm 1$ again. Therefore, there is no such real number $a$ and the Claim follows.

Replace $x$ and $y$ by $-x$ and $-y$ in (1) respectively and compare with (1). We get

$$
f(x+y)^{2}-2 f(x) f(y)=f(-x-y)^{2}-2 f(-x) f(-y) .
$$

Using the Claim, this simplifies to $f(x+y)=f(x)+f(y)+1$. In addition, (5) can be rewritten as $(f(x)+1)^{2}=f\left(x^{2}\right)+1$. Therefore, the function $g$ defined by $g(x)=f(x)+1$ satisfies $g(x+y)=g(x)+g(y)$ and $g(x)^{2}=g\left(x^{2}\right)$. The latter relation shows $g(y)$ is nonnegative for $y \geqslant 0$. For such a function satisfying the Cauchy Equation $g(x+y)=g(x)+g(y)$, it must be monotonic increasing and hence $g(x)=c x$ for some constant $c$.

From $(c x)^{2}=g(x)^{2}=g\left(x^{2}\right)=c x^{2}$, we get $c=0$ or 1 , which corresponds to the two functions $f(x)=-1$ and $f(x)=x-1$ respectively, both of which are solutions to (1) as checked in Solution 1.

Solution 3. As in Solution 2, we find that $f(0)=-1$,

$$
\begin{equation*}
(f(x)+1)^{2}=f\left(x^{2}\right)+1 \tag{8}
\end{equation*}
$$

and

$$
\begin{equation*}
f(x)=f(-x) \quad \text { or } \quad f(x)+f(-x)=-2 \tag{9}
\end{equation*}
$$

for any $x \in \mathbb{R}$. We shall show that one of the statements in (9) holds for all $x \in \mathbb{R}$. Suppose $f(a)=f(-a)$ but $f(a)+f(-a) \neq-2$, while $f(b) \neq f(-b)$ but $f(b)+f(-b)=-2$. Clearly, $a, b \neq 0$ and $f(a), f(b) \neq-1$.

Taking $y=a$ and $y=-a$ in (1) respectively and comparing the two equations obtained, we have $f(x+a)^{2}=f(x-a)^{2}$, that is, $f(x+a)= \pm f(x-a)$. This implies $f(x+2 a)= \pm f(x)$ for all $x \in \mathbb{R}$. Putting $x=b$ and $x=-2 a-b$ respectively, we find $f(2 a+b)= \pm f(b)$ and $f(-2 a-b)= \pm f(-b)= \pm(-2-f(b))$. Since $f(b) \neq-1$, the term $\pm(-2-f(b))$ is distinct from $\pm f(b)$ in any case. So $f(2 a+b) \neq f(-2 a-b)$. From (9), we must have $f(2 a+b)+f(-2 a-b)=-2$. Note that we also have $f(b)+f(-b)=-2$ where $|f(b)|,|f(-b)|$ are equal to $|f(2 a+b)|,|f(-2 a-b)|$ respectively. The only possible case is $f(2 a+b)=f(b)$ and $f(-2 a-b)=f(-b)$.

Applying the argument to $-a$ instead of $a$ and using induction, we have $f(2 k a+b)=f(b)$ and $f(2 k a-b)=f(-b)$ for any integer $k$. Note that $f(b)+f(-b)=-2$ and $f(b) \neq-1$ imply one of $f(b), f(-b)$ is less than -1 . Without loss of generality, assume $f(b)<-1$. We consider $x=\sqrt{2 k a+b}$ in (8) for sufficiently large $k$ so that

$$
(f(x)+1)^{2}=f(2 k a+b)+1=f(b)+1<0
$$

yields a contradiction. Therefore, one of the statements in (9) must hold for all $x \in \mathbb{R}$.

- Case 1. $f(x)=f(-x)$ for any $x \in \mathbb{R}$.

For any $a \in \mathbb{R}$, setting $x=y=\frac{a}{2}$ and $x=-y=\frac{a}{2}$ in (1) respectively and comparing these, we obtain $f(a)^{2}=f(0)^{2}=1$, which means $f(a)= \pm 1$ for all $a \in \mathbb{R}$. If $f(a)=1$ for some $a$, we may assume $a>0$ since $f(a)=f(-a)$. Taking $x=y=\sqrt{a}$ in (1), we get

$$
f(2 \sqrt{a})^{2}=2 f(\sqrt{a})^{2}+\max \{2, f(2 a)\}=2 f(\sqrt{a})^{2}+2 .
$$

Note that the left-hand side is $\pm 1$ while the right-hand side is an even integer. This is a contradiction. Therefore, $f(x)=-1$ for all $x \in \mathbb{R}$, which is clearly a solution.

- Case 2. $f(x)+f(-x)=-2$ for any $x \in \mathbb{R}$.

This case can be handled in the same way as in Solution 2, which yields another solution $f(x)=x-1$.

A8. Determine the largest real number $a$ such that for all $n \geqslant 1$ and for all real numbers $x_{0}, x_{1}, \ldots, x_{n}$ satisfying $0=x_{0}<x_{1}<x_{2}<\cdots<x_{n}$, we have

$$
\begin{equation*}
\frac{1}{x_{1}-x_{0}}+\frac{1}{x_{2}-x_{1}}+\cdots+\frac{1}{x_{n}-x_{n-1}} \geqslant a\left(\frac{2}{x_{1}}+\frac{3}{x_{2}}+\cdots+\frac{n+1}{x_{n}}\right) . \tag{1}
\end{equation*}
$$

Answer. The largest $a$ is $\frac{4}{9}$.
Solution 1. We first show that $a=\frac{4}{9}$ is admissible. For each $2 \leqslant k \leqslant n$, by the CauchySchwarz Inequality, we have

$$
\left(x_{k-1}+\left(x_{k}-x_{k-1}\right)\right)\left(\frac{(k-1)^{2}}{x_{k-1}}+\frac{3^{2}}{x_{k}-x_{k-1}}\right) \geqslant(k-1+3)^{2},
$$

which can be rewritten as

$$
\begin{equation*}
\frac{9}{x_{k}-x_{k-1}} \geqslant \frac{(k+2)^{2}}{x_{k}}-\frac{(k-1)^{2}}{x_{k-1}} . \tag{2}
\end{equation*}
$$

Summing (2) over $k=2,3, \ldots, n$ and adding $\frac{9}{x_{1}}$ to both sides, we have

$$
9 \sum_{k=1}^{n} \frac{1}{x_{k}-x_{k-1}} \geqslant 4 \sum_{k=1}^{n} \frac{k+1}{x_{k}}+\frac{n^{2}}{x_{n}}>4 \sum_{k=1}^{n} \frac{k+1}{x_{k}} .
$$

This shows (1) holds for $a=\frac{4}{9}$.
Next, we show that $a=\frac{4}{9}$ is the optimal choice. Consider the sequence defined by $x_{0}=0$ and $x_{k}=x_{k-1}+k(k+1)$ for $k \geqslant 1$, that is, $x_{k}=\frac{1}{3} k(k+1)(k+2)$. Then the left-hand side of (1) equals

$$
\sum_{k=1}^{n} \frac{1}{k(k+1)}=\sum_{k=1}^{n}\left(\frac{1}{k}-\frac{1}{k+1}\right)=1-\frac{1}{n+1}
$$

while the right-hand side equals

$$
a \sum_{k=1}^{n} \frac{k+1}{x_{k}}=3 a \sum_{k=1}^{n} \frac{1}{k(k+2)}=\frac{3}{2} a \sum_{k=1}^{n}\left(\frac{1}{k}-\frac{1}{k+2}\right)=\frac{3}{2}\left(1+\frac{1}{2}-\frac{1}{n+1}-\frac{1}{n+2}\right) a .
$$

When $n$ tends to infinity, the left-hand side tends to 1 while the right-hand side tends to $\frac{9}{4} a$. Therefore $a$ has to be at most $\frac{4}{9}$.

Hence the largest value of $a$ is $\frac{4}{9}$.
Solution 2. We shall give an alternative method to establish (1) with $a=\frac{4}{9}$. We define $y_{k}=x_{k}-x_{k-1}>0$ for $1 \leqslant k \leqslant n$. By the Cauchy-Schwarz Inequality, for $1 \leqslant k \leqslant n$, we have

$$
\left(y_{1}+y_{2}+\cdots+y_{k}\right)\left(\sum_{j=1}^{k} \frac{1}{y_{j}}\binom{j+1}{2}^{2}\right) \geqslant\left(\binom{2}{2}+\binom{3}{2}+\cdots+\binom{k+1}{2}\right)^{2}=\binom{k+2}{3}^{2}
$$

This can be rewritten as

$$
\begin{equation*}
\frac{k+1}{y_{1}+y_{2}+\cdots+y_{k}} \leqslant \frac{36}{k^{2}(k+1)(k+2)^{2}}\left(\sum_{j=1}^{k} \frac{1}{y_{j}}\binom{j+1}{2}^{2}\right) . \tag{3}
\end{equation*}
$$

Summing (3) over $k=1,2, \ldots, n$, we get

$$
\begin{equation*}
\frac{2}{y_{1}}+\frac{3}{y_{1}+y_{2}}+\cdots+\frac{n+1}{y_{1}+y_{2}+\cdots+y_{n}} \leqslant \frac{c_{1}}{y_{1}}+\frac{c_{2}}{y_{2}}+\cdots+\frac{c_{n}}{y_{n}} \tag{4}
\end{equation*}
$$

where for $1 \leqslant m \leqslant n$,

$$
\begin{aligned}
c_{m} & =36\binom{m+1}{2}^{2} \sum_{k=m}^{n} \frac{1}{k^{2}(k+1)(k+2)^{2}} \\
& =\frac{9 m^{2}(m+1)^{2}}{4} \sum_{k=m}^{n}\left(\frac{1}{k^{2}(k+1)^{2}}-\frac{1}{(k+1)^{2}(k+2)^{2}}\right) \\
& =\frac{9 m^{2}(m+1)^{2}}{4}\left(\frac{1}{m^{2}(m+1)^{2}}-\frac{1}{(n+1)^{2}(n+2)^{2}}\right)<\frac{9}{4} .
\end{aligned}
$$

From (4), the inequality (1) holds for $a=\frac{4}{9}$. This is also the upper bound as can be verified in the same way as Solution 1.

## Combinatorics

C1. The leader of an IMO team chooses positive integers $n$ and $k$ with $n>k$, and announces them to the deputy leader and a contestant. The leader then secretly tells the deputy leader an $n$-digit binary string, and the deputy leader writes down all $n$-digit binary strings which differ from the leader's in exactly $k$ positions. (For example, if $n=3$ and $k=1$, and if the leader chooses 101, the deputy leader would write down 001, 111 and 100.) The contestant is allowed to look at the strings written by the deputy leader and guess the leader's string. What is the minimum number of guesses (in terms of $n$ and $k$ ) needed to guarantee the correct answer?

Answer. The minimum number of guesses is 2 if $n=2 k$ and 1 if $n \neq 2 k$.
Solution 1. Let $X$ be the binary string chosen by the leader and let $X^{\prime}$ be the binary string of length $n$ every digit of which is different from that of $X$. The strings written by the deputy leader are the same as those in the case when the leader's string is $X^{\prime}$ and $k$ is changed to $n-k$. In view of this, we may assume $k \geqslant \frac{n}{2}$. Also, for the particular case $k=\frac{n}{2}$, this argument shows that the strings $X$ and $X^{\prime}$ cannot be distinguished, and hence in that case the contestant has to guess at least twice.

It remains to show that the number of guesses claimed suffices. Consider any string $Y$ which differs from $X$ in $m$ digits where $0<m<2 k$. Without loss of generality, assume the first $m$ digits of $X$ and $Y$ are distinct. Let $Z$ be the binary string obtained from $X$ by changing its first $k$ digits. Then $Z$ is written by the deputy leader. Note that $Z$ differs from $Y$ by $|m-k|$ digits where $|m-k|<k$ since $0<m<2 k$. From this observation, the contestant must know that $Y$ is not the desired string.

As we have assumed $k \geqslant \frac{n}{2}$, when $n<2 k$, every string $Y \neq X$ differs from $X$ in fewer than $2 k$ digits. When $n=2 k$, every string except $X$ and $X^{\prime}$ differs from $X$ in fewer than $2 k$ digits. Hence, the answer is as claimed.

Solution 2. Firstly, assume $n \neq 2 k$. Without loss of generality suppose the first digit of the leader's string is 1 . Then among the $\binom{n}{k}$ strings written by the deputy leader, $\binom{n-1}{k}$ will begin with 1 and $\binom{n-1}{k-1}$ will begin with 0 . Since $n \neq 2 k$, we have $k+(k-1) \neq n-1$ and so $\binom{n-1}{k} \neq\binom{ n-1}{k-1}$. Thus, by counting the number of strings written by the deputy leader that start with 0 and 1 , the contestant can tell the first digit of the leader's string. The same can be done on the other digits, so 1 guess suffices when $n \neq 2 k$.

Secondly, for the case $n=2$ and $k=1$, the answer is clearly 2 . For the remaining cases where $n=2 k>2$, the deputy leader would write down the same strings if the leader's string $X$ is replaced by $X^{\prime}$ obtained by changing each digit of $X$. This shows at least 2 guesses are needed. We shall show that 2 guesses suffice in this case. Suppose the first two digits of the leader's string are the same. Then among the strings written by the deputy leader, the prefices 01 and 10 will occur $\binom{2 k-2}{k-1}$ times each, while the prefices 00 and 11 will occur $\binom{2 k-2}{k}$ times each. The two numbers are interchanged if the first two digits of the leader's string are different. Since $\binom{2 k-2}{k-1} \neq\binom{ 2 k-2}{k}$, the contestant can tell whether the first two digits of the leader's string are the same or not. He can work out the relation of the first digit and the
other digits in the same way and reduce the leader's string to only 2 possibilities. The proof is complete.

C2. Find all positive integers $n$ for which all positive divisors of $n$ can be put into the cells of a rectangular table under the following constraints:

- each cell contains a distinct divisor;
- the sums of all rows are equal; and
- the sums of all columns are equal.

Answer. 1.
Solution 1. Suppose all positive divisors of $n$ can be arranged into a rectangular table of size $k \times l$ where the number of rows $k$ does not exceed the number of columns $l$. Let the sum of numbers in each column be $s$. Since $n$ belongs to one of the columns, we have $s \geqslant n$, where equality holds only when $n=1$.

For $j=1,2, \ldots, l$, let $d_{j}$ be the largest number in the $j$-th column. Without loss of generality, assume $d_{1}>d_{2}>\cdots>d_{l}$. Since these are divisors of $n$, we have

$$
\begin{equation*}
d_{l} \leqslant \frac{n}{l} \tag{1}
\end{equation*}
$$

As $d_{l}$ is the maximum entry of the $l$-th column, we must have

$$
\begin{equation*}
d_{l} \geqslant \frac{s}{k} \geqslant \frac{n}{k} . \tag{2}
\end{equation*}
$$

The relations (1) and (2) combine to give $\frac{n}{l} \geqslant \frac{n}{k}$, that is, $k \geqslant l$. Together with $k \leqslant l$, we conclude that $k=l$. Then all inequalities in (1) and (2) are equalities. In particular, $s=n$ and so $n=1$, in which case the conditions are clearly satisfied.

Solution 2. Clearly $n=1$ works. Then we assume $n>1$ and let its prime factorization be $n=p_{1}^{r_{1}} p_{2}^{r_{2}} \cdots p_{t}^{r_{t}}$. Suppose the table has $k$ rows and $l$ columns with $1<k \leqslant l$. Note that $k l$ is the number of positive divisors of $n$ and the sum of all entries is the sum of positive divisors of $n$, which we denote by $\sigma(n)$. Consider the column containing $n$. Since the column sum is $\frac{\sigma(n)}{l}$, we must have $\frac{\sigma(n)}{l}>n$. Therefore, we have

$$
\begin{aligned}
\left(r_{1}+1\right)\left(r_{2}+1\right) \cdots\left(r_{t}+1\right) & =k l \leqslant l^{2}<\left(\frac{\sigma(n)}{n}\right)^{2} \\
& =\left(1+\frac{1}{p_{1}}+\cdots+\frac{1}{p_{1}^{r_{1}}}\right)^{2} \cdots\left(1+\frac{1}{p_{t}}+\cdots+\frac{1}{p_{t}^{r_{t}}}\right)^{2} .
\end{aligned}
$$

This can be rewritten as

$$
\begin{equation*}
f\left(p_{1}, r_{1}\right) f\left(p_{2}, r_{2}\right) \cdots f\left(p_{t}, r_{t}\right)<1 \tag{3}
\end{equation*}
$$

where

$$
f(p, r)=\frac{r+1}{\left(1+\frac{1}{p}+\cdots+\frac{1}{p^{r}}\right)^{2}}=\frac{(r+1)\left(1-\frac{1}{p}\right)^{2}}{\left(1-\frac{1}{p^{r+1}}\right)^{2}}
$$

Direct computation yields

$$
f(2,1)=\frac{8}{9}, \quad f(2,2)=\frac{48}{49}, \quad f(3,1)=\frac{9}{8} .
$$

Also, we find that

$$
\begin{aligned}
& f(2, r) \geqslant\left(1-\frac{1}{2^{r+1}}\right)^{-2}>1 \quad \text { for } r \geqslant 3 \\
& f(3, r) \geqslant \frac{4}{3}\left(1-\frac{1}{3^{r+1}}\right)^{-2}>\frac{4}{3}>\frac{9}{8} \quad \text { for } r \geqslant 2, \text { and } \\
& f(p, r) \geqslant \frac{32}{25}\left(1-\frac{1}{p^{r+1}}\right)^{-2}>\frac{32}{25}>\frac{9}{8} \quad \text { for } p \geqslant 5 .
\end{aligned}
$$

From these values and bounds, it is clear that (3) holds only when $n=2$ or 4 . In both cases, it is easy to see that the conditions are not satisfied. Hence, the only possible $n$ is 1 .

C3. Let $n$ be a positive integer relatively prime to 6 . We paint the vertices of a regular $n$-gon with three colours so that there is an odd number of vertices of each colour. Show that there exists an isosceles triangle whose three vertices are of different colours.

Solution. For $k=1,2,3$, let $a_{k}$ be the number of isosceles triangles whose vertices contain exactly $k$ colours. Suppose on the contrary that $a_{3}=0$. Let $b, c, d$ be the number of vertices of the three different colours respectively. We now count the number of pairs $(\triangle, E)$ where $\triangle$ is an isosceles triangle and $E$ is a side of $\triangle$ whose endpoints are of different colours.

On the one hand, since we have assumed $a_{3}=0$, each triangle in the pair must contain exactly two colours, and hence each triangle contributes twice. Thus the number of pairs is $2 a_{2}$.

On the other hand, if we pick any two vertices $A, B$ of distinct colours, then there are three isosceles triangles having these as vertices, two when $A B$ is not the base and one when $A B$ is the base since $n$ is odd. Note that the three triangles are all distinct as $(n, 3)=1$. In this way, we count the number of pairs to be $3(b c+c d+d b)$. However, note that $2 a_{2}$ is even while $3(b c+c d+d b)$ is odd, as each of $b, c, d$ is. This yields a contradiction and hence $a_{3} \geqslant 1$.

Comment. A slightly stronger version of this problem is to replace the condition $(n, 6)=1$ by $n$ being odd (where equilateral triangles are regarded as isosceles triangles). In that case, the only difference in the proof is that by fixing any two vertices $A, B$, one can find exactly one or three isosceles triangles having these as vertices. But since only parity is concerned in the solution, the proof goes the same way.

The condition that there is an odd number of vertices of each colour is necessary, as can be seen from the following example. Consider $n=25$ and we label the vertices $A_{0}, A_{1}, \ldots, A_{24}$. Suppose colour 1 is used for $A_{0}$, colour 2 is used for $A_{5}, A_{10}, A_{15}, A_{20}$, while colour 3 is used for the remaining vertices. Then any isosceles triangle having colours 1 and 2 must contain $A_{0}$ and one of $A_{5}, A_{10}, A_{15}, A_{20}$. Clearly, the third vertex must have index which is a multiple of 5 so it is not of colour 3 .

C4. Find all positive integers $n$ for which we can fill in the entries of an $n \times n$ table with the following properties:

- each entry can be one of $I, M$ and $O$;
- in each row and each column, the letters $I, M$ and $O$ occur the same number of times; and
- in any diagonal whose number of entries is a multiple of three, the letters $I, M$ and $O$ occur the same number of times.

Answer. $n$ can be any multiple of 9 .
Solution. We first show that such a table exists when $n$ is a multiple of 9 . Consider the following $9 \times 9$ table.

$$
\left(\begin{array}{ccccccccc}
I & I & I & M & M & M & O & O & O  \tag{1}\\
M & M & M & O & O & O & I & I & I \\
O & O & O & I & I & I & M & M & M \\
I & I & I & M & M & M & O & O & O \\
M & M & M & O & O & O & I & I & I \\
O & O & O & I & I & I & M & M & M \\
I & I & I & M & M & M & O & O & O \\
M & M & M & O & O & O & I & I & I \\
O & O & O & I & I & I & M & M & M
\end{array}\right)
$$

It is a direct checking that the table (1) satisfies the requirements. For $n=9 k$ where $k$ is a positive integer, we form an $n \times n$ table using $k \times k$ copies of (1). For each row and each column of the table of size $n$, since there are three $I$ 's, three $M$ 's and three $O$ 's for any nine consecutive entries, the numbers of $I, M$ and $O$ are equal. In addition, every diagonal of the large table whose number of entries is divisible by 3 intersects each copy of (1) at a diagonal with number of entries divisible by 3 (possibly zero). Therefore, every such diagonal also contains the same number of $I, M$ and $O$.

Next, consider any $n \times n$ table for which the requirements can be met. As the number of entries of each row should be a multiple of 3 , we let $n=3 k$ where $k$ is a positive integer. We divide the whole table into $k \times k$ copies of $3 \times 3$ blocks. We call the entry at the centre of such a $3 \times 3$ square a vital entry. We also call any row, column or diagonal that contains at least one vital entry a vital line. We compute the number of pairs $(l, c)$ where $l$ is a vital line and $c$ is an entry belonging to $l$ that contains the letter $M$. We let this number be $N$.

On the one hand, since each vital line contains the same number of $I, M$ and $O$, it is obvious that each vital row and each vital column contain $k$ occurrences of $M$. For vital diagonals in either direction, we count there are exactly

$$
1+2+\cdots+(k-1)+k+(k-1)+\cdots+2+1=k^{2}
$$

occurrences of $M$. Therefore, we have $N=4 k^{2}$.

On the other hand, there are $3 k^{2}$ occurrences of $M$ in the whole table. Note that each entry belongs to exactly 1 or 4 vital lines. Therefore, $N$ must be congruent to $3 k^{2} \bmod 3$.

From the double counting, we get $4 k^{2} \equiv 3 k^{2}(\bmod 3)$, which forces $k$ to be a multiple of 3. Therefore, $n$ has to be a multiple of 9 and the proof is complete.

C5. Let $n \geqslant 3$ be a positive integer. Find the maximum number of diagonals of a regular $n$-gon one can select, so that any two of them do not intersect in the interior or they are perpendicular to each other.

Answer. $n-2$ if $n$ is even and $n-3$ if $n$ is odd.
Solution 1. We consider two cases according to the parity of $n$.

- Case 1. $n$ is odd.

We first claim that no pair of diagonals is perpendicular. Suppose $A, B, C, D$ are vertices where $A B$ and $C D$ are perpendicular, and let $E$ be the vertex lying on the perpendicular bisector of $A B$. Let $E^{\prime}$ be the opposite point of $E$ on the circumcircle of the regular polygon. Since $E C=E^{\prime} D$ and $C, D, E$ are vertices of the regular polygon, $E^{\prime}$ should also belong to the polygon. This contradicts the fact that a regular polygon with an odd number of vertices does not contain opposite points on the circumcircle.


Therefore in the odd case we can only select diagonals which do not intersect. In the maximal case these diagonals should divide the regular $n$-gon into $n-2$ triangles, so we can select at most $n-3$ diagonals. This can be done, for example, by selecting all diagonals emanated from a particular vertex.

- Case 2. $n$ is even.

If there is no intersection, then the proof in the odd case works. Suppose there are two perpendicular diagonals selected. We consider the set $S$ of all selected diagonals parallel to one of them which intersect with some selected diagonals. Suppose $S$ contains $k$ diagonals and the number of distinct endpoints of the $k$ diagonals is $l$.

Firstly, consider the longest diagonal in one of the two directions in $S$. No other diagonal in $S$ can start from either endpoint of that diagonal, since otherwise it has to meet another longer diagonal in $S$. The same holds true for the other direction. Ignoring these two longest diagonals and their four endpoints, the remaining $k-2$ diagonals share $l-4$ endpoints where each endpoint can belong to at most two diagonals. This gives $2(l-4) \geqslant 2(k-2)$, so that $k \leqslant l-2$.


Consider a group of consecutive vertices of the regular $n$-gon so that each of the two outermost vertices is an endpoint of a diagonal in $S$, while the interior points are not. There are $l$ such groups. We label these groups $P_{1}, P_{2}, \ldots, P_{l}$ in this order. We claim that each selected diagonal outside $S$ must connect vertices of the same group $P_{i}$. Consider any diagonal $d$ joining vertices from distinct groups $P_{i}$ and $P_{j}$. Let $d_{1}$ and $d_{2}$ be two diagonals in $S$ each having one of the outermost points of $P_{i}$ as endpoint. Then $d$ must meet either $d_{1}, d_{2}$ or a diagonal in $S$ which is perpendicular to both $d_{1}$ and $d_{2}$. In any case $d$ should belong to $S$ by definition, which is a contradiction.

Within the same group $P_{i}$, there are no perpendicular diagonals since the vertices belong to the same side of a diameter of the circumcircle. Hence there can be at most $\left|P_{i}\right|-2$ selected diagonals within $P_{i}$, including the one joining the two outermost points of $P_{i}$ when $\left|P_{i}\right|>2$. Therefore, the maximum number of diagonals selected is

$$
\sum_{i=1}^{l}\left(\left|P_{i}\right|-2\right)+k=\sum_{i=1}^{l}\left|P_{i}\right|-2 l+k=(n+l)-2 l+k=n-l+k \leqslant n-2 .
$$

This upper bound can be attained as follows. We take any vertex $A$ and let $A^{\prime}$ be the vertex for which $A A^{\prime}$ is a diameter of the circumcircle. If we select all diagonals emanated from $A$ together with the diagonal $d^{\prime}$ joining the two neighbouring vertices of $A^{\prime}$, then the only pair of diagonals that meet each other is $A A^{\prime}$ and $d^{\prime}$, which are perpendicular to each other. In total we can take $n-2$ diagonals.


Solution 2. The constructions and the odd case are the same as Solution 1. Instead of dealing separately with the case where $n$ is even, we shall prove by induction more generally that we can select at most $n-2$ diagonals for any cyclic $n$-gon with circumcircle $\Gamma$.

The base case $n=3$ is trivial since there is no diagonal at all. Suppose the upper bound holds for any cyclic polygon with fewer than $n$ sides. For a cyclic $n$-gon, if there is a selected diagonal which does not intersect any other selected diagonal, then this diagonal divides the $n$-gon into an $m$-gon and an $l$-gon (with $m+l=n+2$ ) so that each selected diagonal belongs to one of them. Without loss of generality, we may assume the $m$-gon lies on the same side of a diameter of $\Gamma$. Then no two selected diagonals of the $m$-gon can intersect, and hence we can select at most $m-3$ diagonals. Also, we can apply the inductive hypothesis to the $l$-gon. This shows the maximum number of selected diagonals is $(m-3)+(l-2)+1=n-2$.

It remains to consider the case when all selected diagonals meet at least one other selected diagonal. Consider a pair of selected perpendicular diagonals $d_{1}, d_{2}$. They divide the circumference of $\Gamma$ into four arcs, each of which lies on the same side of a diameter of $\Gamma$. If there are two selected diagonals intersecting each other and neither is parallel to $d_{1}$ or $d_{2}$, then their endpoints must belong to the same arc determined by $d_{1}, d_{2}$, and hence they cannot be perpendicular. This violates the condition, and hence all selected diagonals must have the same direction as one of $d_{1}, d_{2}$.


Take the longest selected diagonal in one of the two directions. We argue as in Solution 1 that its endpoints do not belong to any other selected diagonal. The same holds true for the longest diagonal in the other direction. Apart from these four endpoints, each of the remaining $n-4$ vertices can belong to at most two selected diagonals. Thus we can select at most $\frac{1}{2}(2(n-4)+4)=n-2$ diagonals. Then the proof follows by induction.

C6. There are $n \geqslant 3$ islands in a city. Initially, the ferry company offers some routes between some pairs of islands so that it is impossible to divide the islands into two groups such that no two islands in different groups are connected by a ferry route.

After each year, the ferry company will close a ferry route between some two islands $X$ and $Y$. At the same time, in order to maintain its service, the company will open new routes according to the following rule: for any island which is connected by a ferry route to exactly one of $X$ and $Y$, a new route between this island and the other of $X$ and $Y$ is added.

Suppose at any moment, if we partition all islands into two nonempty groups in any way, then it is known that the ferry company will close a certain route connecting two islands from the two groups after some years. Prove that after some years there will be an island which is connected to all other islands by ferry routes.

Solution. Initially, we pick any pair of islands $A$ and $B$ which are connected by a ferry route and put $A$ in set $\mathcal{A}$ and $B$ in set $\mathcal{B}$. From the condition, without loss of generality there must be another island which is connected to $A$. We put such an island $C$ in set $\mathcal{B}$. We say that two sets of islands form a network if each island in one set is connected to each island in the other set.

Next, we shall included all islands to $\mathcal{A} \cup \mathcal{B}$ one by one. Suppose we have two sets $\mathcal{A}$ and $\mathcal{B}$ which form a network where $3 \leqslant|\mathcal{A} \cup \mathcal{B}|<n$. This relation no longer holds only when a ferry route between islands $A \in \mathcal{A}$ and $B \in \mathcal{B}$ is closed. In that case, we define $\mathcal{A}^{\prime}=\{A, B\}$, and $\mathcal{B}^{\prime}=(\mathcal{A} \cup \mathcal{B})-\{A, B\}$. Note that $\mathcal{B}^{\prime}$ is nonempty. Consider any island $C \in \mathcal{A}-\{A\}$. From the relation of $\mathcal{A}$ and $\mathcal{B}$, we know that $C$ is connected to $B$. If $C$ was not connected to $A$ before the route between $A$ and $B$ closes, then there will be a route added between $C$ and $A$ afterwards. Hence, $C$ must now be connected to both $A$ and $B$. The same holds true for any island in $\mathcal{B}-\{B\}$. Therefore, $\mathcal{A}^{\prime}$ and $\mathcal{B}^{\prime}$ form a network, and $\mathcal{A}^{\prime} \cup \mathcal{B}^{\prime}=\mathcal{A} \cup \mathcal{B}$. Hence these islands can always be partitioned into sets $\mathcal{A}$ and $\mathcal{B}$ which form a network.

As $|\mathcal{A} \cup \mathcal{B}|<n$, there are some islands which are not included in $\mathcal{A} \cup \mathcal{B}$. From the condition, after some years there must be a ferry route between an island $A$ in $\mathcal{A} \cup \mathcal{B}$ and an island $D$ outside $\mathcal{A} \cup \mathcal{B}$ which closes. Without loss of generality assume $A \in \mathcal{A}$. Then each island in $\mathcal{B}$ must then be connected to $D$, no matter it was or not before. Hence, we can put $D$ in set $\mathcal{A}$ so that the new sets $\mathcal{A}$ and $\mathcal{B}$ still form a network and the size of $\mathcal{A} \cup \mathcal{B}$ is increased by 1 . The same process can be done to increase the size of $\mathcal{A} \cup \mathcal{B}$. Eventually, all islands are included in this way so we may now assume $|\mathcal{A} \cup \mathcal{B}|=n$.

Suppose a ferry route between $A \in \mathcal{A}$ and $B \in \mathcal{B}$ is closed after some years. We put $A$ and $B$ in set $\mathcal{A}^{\prime}$ and all remaining islands in set $\mathcal{B}^{\prime}$. Then $\mathcal{A}^{\prime}$ and $\mathcal{B}^{\prime}$ form a network. This relation no longer holds only when a route between $A$, without loss of generality, and $C \in \mathcal{B}^{\prime}$ is closed. Since this must eventually occur, at that time island $B$ will be connected to all other islands and the result follows.

C7. Let $n \geqslant 2$ be an integer. In the plane, there are $n$ segments given in such a way that any two segments have an intersection point in the interior, and no three segments intersect at a single point. Jeff places a snail at one of the endpoints of each of the segments and claps his hands $n-1$ times. Each time when he claps his hands, all the snails move along their own segments and stay at the next intersection points until the next clap. Since there are $n-1$ intersection points on each segment, all snails will reach the furthest intersection points from their starting points after $n-1$ claps.
(a) Prove that if $n$ is odd then Jeff can always place the snails so that no two of them ever occupy the same intersection point.
(b) Prove that if $n$ is even then there must be a moment when some two snails occupy the same intersection point no matter how Jeff places the snails.

Solution. We consider a big disk which contains all the segments. We extend each segment to a line $l_{i}$ so that each of them cuts the disk at two distinct points $A_{i}, B_{i}$.
(a) For odd $n$, we travel along the circumference of the disk and mark each of the points $A_{i}$ or $B_{i}$ 'in' and 'out' alternately. Since every pair of lines intersect in the disk, there are exactly $n-1$ points between $A_{i}$ and $B_{i}$ for any fixed $1 \leqslant i \leqslant n$. As $n$ is odd, this means one of $A_{i}$ and $B_{i}$ is marked 'in' and the other is marked 'out'. Then Jeff can put a snail on the endpoint of each segment which is closer to the 'in' side of the corresponding line. We claim that the snails on $l_{i}$ and $l_{j}$ do not meet for any pairs $i, j$, hence proving part (a).


Without loss of generality, we may assume the snails start at $A_{i}$ and $A_{j}$ respectively. Let $l_{i}$ intersect $l_{j}$ at $P$. Note that there is an odd number of points between arc $A_{i} A_{j}$. Each of these points belongs to a line $l_{k}$. Such a line $l_{k}$ must intersect exactly one of
the segments $A_{i} P$ and $A_{j} P$, making an odd number of intersections. For the other lines, they may intersect both segments $A_{i} P$ and $A_{j} P$, or meet none of them. Therefore, the total number of intersection points on segments $A_{i} P$ and $A_{j} P$ (not counting $P$ ) is odd. However, if the snails arrive at $P$ at the same time, then there should be the same number of intersections on $A_{i} P$ and $A_{j} P$, which gives an even number of intersections. This is a contradiction so the snails do not meet each other.
(b) For even $n$, we consider any way that Jeff places the snails and mark each of the points $A_{i}$ or $B_{i}$ 'in' and 'out' according to the directions travelled by the snails. In this case there must be two neighbouring points $A_{i}$ and $A_{j}$ both of which are marked 'in'. Let $P$ be the intersection of the segments $A_{i} B_{i}$ and $A_{j} B_{j}$. Then any other segment meeting one of the segments $A_{i} P$ and $A_{j} P$ must also meet the other one, and so the number of intersections on $A_{i} P$ and $A_{j} P$ are the same. This shows the snails starting from $A_{i}$ and $A_{j}$ will meet at $P$.

Comment. The conclusions do not hold for pseudosegments, as can be seen from the following examples.


C8. Let $n$ be a positive integer. Determine the smallest positive integer $k$ with the following property: it is possible to mark $k$ cells on a $2 n \times 2 n$ board so that there exists a unique partition of the board into $1 \times 2$ and $2 \times 1$ dominoes, none of which contains two marked cells.

Answer. $2 n$.
Solution. We first construct an example of marking $2 n$ cells satisfying the requirement. Label the rows and columns $1,2, \ldots, 2 n$ and label the cell in the $i$-th row and the $j$-th column $(i, j)$.

For $i=1,2, \ldots, n$, we mark the cells $(i, i)$ and $(i, i+1)$. We claim that the required partition exists and is unique. The two diagonals of the board divide the board into four regions. Note that the domino covering cell $(1,1)$ must be vertical. This in turn shows that each domino covering $(2,2),(3,3), \ldots,(n, n)$ is vertical. By induction, the dominoes in the left region are all vertical. By rotational symmetry, the dominoes in the bottom region are horizontal, and so on. This shows that the partition exists and is unique.


It remains to show that this value of $k$ is the smallest possible. Assume that only $k<2 n$ cells are marked, and there exists a partition $P$ satisfying the requirement. It suffices to show there exists another desirable partition distinct from $P$. Let $d$ be the main diagonal of the board.

Construct the following graph with edges of two colours. Its vertices are the cells of the board. Connect two vertices with a red edge if they belong to the same domino of $P$. Connect two vertices with a blue edge if their reflections in $d$ are connected by a red edge. It is possible that two vertices are connected by edges of both colours. Clearly, each vertex has both red and blue degrees equal to 1 . Thus the graph splits into cycles where the colours of edges in each cycle alternate (a cycle may have length 2).

Consider any cell $c$ lying on the diagonal $d$. Its two edges are symmetrical with respect to $d$. Thus they connect $c$ to different cells. This shows $c$ belongs to a cycle $C(c)$ of length at least 4. Consider a part of this cycle $c_{0}, c_{1}, \ldots, c_{m}$ where $c_{0}=c$ and $m$ is the least positive integer such that $c_{m}$ lies on $d$. Clearly, $c_{m}$ is distinct from $c$. From the construction, the path symmetrical to this with respect to $d$ also lies in the graph, and so these paths together form $C(c)$. Hence, $C(c)$ contains exactly two cells from $d$. Then all $2 n$ cells in $d$ belong to $n$ cycles $C_{1}, C_{2}, \ldots, C_{n}$, each has length at least 4.

By the Pigeonhole Principle, there exists a cycle $C_{i}$ containing at most one of the $k$ marked cells. We modify $P$ as follows. We remove all dominoes containing the vertices of $C_{i}$, which
correspond to the red edges of $C_{i}$. Then we put the dominoes corresponding to the blue edges of $C_{i}$. Since $C_{i}$ has at least 4 vertices, the resultant partition $P^{\prime}$ is different from $P$. Clearly, no domino in $P^{\prime}$ contains two marked cells as $C_{i}$ contains at most one marked cell. This shows the partition is not unique and hence $k$ cannot be less than $2 n$.

## Geometry

G1. In a convex pentagon $A B C D E$, let $F$ be a point on $A C$ such that $\angle F B C=90^{\circ}$. Suppose triangles $A B F, A C D$ and $A D E$ are similar isosceles triangles with

$$
\begin{equation*}
\angle F A B=\angle F B A=\angle D A C=\angle D C A=\angle E A D=\angle E D A . \tag{1}
\end{equation*}
$$

Let $M$ be the midpoint of $C F$. Point $X$ is chosen such that $A M X E$ is a parallelogram. Show that $B D, E M$ and $F X$ are concurrent.

Solution 1. Denote the common angle in (1) by $\theta$. As $\triangle A B F \sim \triangle A C D$, we have $\frac{A B}{A C}=\frac{A F}{A D}$ so that $\triangle A B C \sim \triangle A F D$. From $E A=E D$, we get

$$
\angle A F D=\angle A B C=90^{\circ}+\theta=180^{\circ}-\frac{1}{2} \angle A E D
$$

Hence, $F$ lies on the circle with centre $E$ and radius $E A$. In particular, $E F=E A=E D$. As $\angle E F A=\angle E A F=2 \theta=\angle B F C$, points $B, F, E$ are collinear.

As $\angle E D A=\angle M A D$, we have $E D / / A M$ and hence $E, D, X$ are collinear. As $M$ is the midpoint of $C F$ and $\angle C B F=90^{\circ}$, we get $M F=M B$. In the isosceles triangles $E F A$ and $M F B$, we have $\angle E F A=\angle M F B$ and $A F=B F$. Therefore, they are congruent to each other. Then we have $B M=A E=X M$ and $B E=B F+F E=A F+F M=A M=E X$. This shows $\triangle E M B \cong \triangle E M X$. As $F$ and $D$ lie on $E B$ and $E X$ respectively and $E F=E D$, we know that lines $B D$ and $X F$ are symmetric with respect to $E M$. It follows that the three lines are concurrent.


Solution 2. From $\angle C A D=\angle E D A$, we have $A C / / E D$. Together with $A C / / E X$, we know that $E, D, X$ are collinear. Denote the common angle in (1) by $\theta$. From $\triangle A B F \sim \triangle A C D$, we get $\frac{A B}{A C}=\frac{A F}{A D}$ so that $\triangle A B C \sim \triangle A F D$. This yields $\angle A F D=\angle A B C=90^{\circ}+\theta$ and hence $\angle F D C=90^{\circ}$, implying that $B C D F$ is cyclic. Let $\Gamma_{1}$ be its circumcircle.

Next, from $\triangle A B F \sim \triangle A D E$, we have $\frac{A B}{A D}=\frac{A F}{A E}$ so that $\triangle A B D \sim \triangle A F E$. Therefore,

$$
\angle A F E=\angle A B D=\theta+\angle F B D=\theta+\angle F C D=2 \theta=180^{\circ}-\angle B F A .
$$

This implies $B, F, E$ are collinear. Note that $F$ is the incentre of triangle $D A B$. Point $E$ lies on the internal angle bisector of $\angle D B A$ and lies on the perpendicular bisector of $A D$. It follows that $E$ lies on the circumcircle $\Gamma_{2}$ of triangle $A B D$, and $E A=E F=E D$.

Also, since $C F$ is a diameter of $\Gamma_{1}$ and $M$ is the midpoint of $C F, M$ is the centre of $\Gamma_{1}$ and hence $\angle A M D=2 \theta=\angle A B D$. This shows $M$ lies on $\Gamma_{2}$. Next, $\angle M D X=\angle M A E=\angle D X M$ since $A M X E$ is a parallelogram. Hence $M D=M X$ and $X$ lies on $\Gamma_{1}$.


We now have two ways to complete the solution.

- Method 1. From $E F=E A=X M$ and $E X / / F M, E F M X$ is an isosceles trapezoid and is cyclic. Denote its circumcircle by $\Gamma_{3}$. Since $B D, E M, F X$ are the three radical axes of $\Gamma_{1}, \Gamma_{2}, \Gamma_{3}$, they must be concurrent.
- Method 2. As $\angle D M F=2 \theta=\angle B F M$, we have $D M / / E B$. Also,

$$
\angle B F D+\angle X B F=\angle B F C+\angle C F D+90^{\circ}-\angle C B X=2 \theta+\left(90^{\circ}-\theta\right)+90^{\circ}-\theta=180^{\circ}
$$

implies $D F / / X B$. These show the corresponding sides of triangles $D M F$ and $B E X$ are parallel. By Desargues' Theorem, the two triangles are perspective and hence $D B, M E, F X$ meet at a point.

Comment. In Solution 2, both the Radical Axis Theorem and Desargues' Theorem could imply $D B, M E, F X$ are parallel. However, this is impossible as can be seen from the configuration. For example, it is obvious that $D B$ and $M E$ meet each other.

Solution 3. Let the common angle in (1) be $\theta$. From $\triangle A B F \sim \triangle A C D$, we have $\frac{A B}{A C}=\frac{A F}{A D}$ so that $\triangle A B C \sim \triangle A F D$. Then $\angle A D F=\angle A C B=90^{\circ}-2 \theta=90^{\circ}-\angle B A D$ and hence $D F \perp A B$. As $F A=F B$, this implies $\triangle D A B$ is isosceles with $D A=D B$. Then $F$ is the incentre of $\triangle D A B$.

Next, from $\angle A E D=180^{\circ}-2 \theta=180^{\circ}-\angle D B A$, points $A, B, D, E$ are concyclic. Since we also have $E A=E D$, this shows $E, F, B$ are collinear and $E A=E F=E D$.


Note that $C$ lies on the internal angle bisector of $\angle B A D$ and lies on the external angle bisector of $\angle D B A$. It follows that it is the $A$-excentre of triangle $D A B$. As $M$ is the midpoint of $C F, M$ lies on the circumcircle of triangle $D A B$ and it is the centre of the circle passing through $D, F, B, C$. By symmetry, $D E F M$ is a rhombus. Then the midpoints of $A X, E M$ and $D F$ coincide, and it follows that $D A F X$ is a parallelogram.

Let $P$ be the intersection of $B D$ and $E M$, and $Q$ be the intersection of $A D$ and $B E$. From $\angle B A C=\angle D C A$, we know that $D C, A B, E M$ are parallel. Thus we have $\frac{D P}{P B}=\frac{C M}{M A}$. This is further equal to $\frac{A E}{B E}$ since $C M=D M=D E=A E$ and $M A=B E$. From $\triangle A E Q \sim \triangle B E A$, we find that

$$
\frac{D P}{P B}=\frac{A E}{B E}=\frac{A Q}{B A}=\frac{Q F}{F B}
$$

by the Angle Bisector Theorem. This implies $Q D / / F P$ and hence $F, P, X$ are collinear, as desired.

G2. Let $A B C$ be a triangle with circumcircle $\Gamma$ and incentre $I$. Let $M$ be the midpoint of side $B C$. Denote by $D$ the foot of perpendicular from $I$ to side $B C$. The line through $I$ perpendicular to $A I$ meets sides $A B$ and $A C$ at $F$ and $E$ respectively. Suppose the circumcircle of triangle $A E F$ intersects $\Gamma$ at a point $X$ other than $A$. Prove that lines $X D$ and $A M$ meet on $\Gamma$.

Solution 1. Let $A M$ meet $\Gamma$ again at $Y$ and $X Y$ meet $B C$ at $D^{\prime}$. It suffices to show $D^{\prime}=D$. We shall apply the following fact.

- Claim. For any cyclic quadrilateral $P Q R S$ whose diagonals meet at $T$, we have

$$
\frac{Q T}{T S}=\frac{P Q \cdot Q R}{P S \cdot S R}
$$

Proof. We use $\left[W_{1} W_{2} W_{3}\right]$ to denote the area of $W_{1} W_{2} W_{3}$. Then

$$
\frac{Q T}{T S}=\frac{[P Q R]}{[P S R]}=\frac{\frac{1}{2} P Q \cdot Q R \sin \angle P Q R}{\frac{1}{2} P S \cdot S R \sin \angle P S R}=\frac{P Q \cdot Q R}{P S \cdot S R}
$$

Applying the Claim to $A B Y C$ and $X B Y C$ respectively, we have $1=\frac{B M}{M C}=\frac{A B \cdot B Y}{A C \cdot C Y}$ and $\frac{B D^{\prime}}{D^{\prime} C}=\frac{X B \cdot B Y}{X C \cdot C Y}$. These combine to give

$$
\begin{equation*}
\frac{B D^{\prime}}{C D^{\prime}}=\frac{X B}{X C} \cdot \frac{B Y}{C Y}=\frac{X B}{X C} \cdot \frac{A C}{A B} \tag{1}
\end{equation*}
$$

Next, we use directed angles to find that $\measuredangle X B F=\measuredangle X B A=\measuredangle X C A=\measuredangle X C E$ and $\measuredangle X F B=\measuredangle X F A=\measuredangle X E A=\measuredangle X E C$. This shows triangles $X B F$ and $X C E$ are directly similar. In particular, we have

$$
\begin{equation*}
\frac{X B}{X C}=\frac{B F}{C E} . \tag{2}
\end{equation*}
$$

In the following, we give two ways to continue the proof.

- Method 1. Here is a geometrical method. As $\angle F I B=\angle A I B-90^{\circ}=\frac{1}{2} \angle A C B=\angle I C B$ and $\angle F B I=\angle I B C$, the triangles $F B I$ and $I B C$ are similar. Analogously, triangles $E I C$ and $I B C$ are also similar. Hence, we get

$$
\begin{equation*}
\frac{F B}{I B}=\frac{B I}{B C} \quad \text { and } \quad \frac{E C}{I C}=\frac{I C}{B C} \tag{3}
\end{equation*}
$$



Next, construct a line parallel to $B C$ and tangent to the incircle. Suppose it meets sides $A B$ and $A C$ at $B_{1}$ and $C_{1}$ respectively. Let the incircle touch $A B$ and $A C$ at $B_{2}$ and $C_{2}$ respectively. By homothety, the line $B_{1} I$ is parallel to the external angle bisector of $\angle A B C$, and hence $\angle B_{1} I B=90^{\circ}$. Since $\angle B B_{2} I=90^{\circ}$, we get $B B_{2} \cdot B B_{1}=B I^{2}$, and similarly $C C_{2} \cdot C C_{1}=C I^{2}$. Hence,

$$
\begin{equation*}
\frac{B I^{2}}{C I^{2}}=\frac{B B_{2} \cdot B B_{1}}{C C_{2} \cdot C C_{1}}=\frac{B B_{1}}{C C_{1}} \cdot \frac{B D}{C D}=\frac{A B}{A C} \cdot \frac{B D}{C D} \tag{4}
\end{equation*}
$$

Combining (1), (2), (3) and (4), we conclude

$$
\frac{B D^{\prime}}{C D^{\prime}}=\frac{X B}{X C} \cdot \frac{A C}{A B}=\frac{B F}{C E} \cdot \frac{A C}{A B}=\frac{B I^{2}}{C I^{2}} \cdot \frac{A C}{A B}=\frac{B D}{C D}
$$

so that $D^{\prime}=D$. The result then follows.

- Method 2. We continue the proof of Solution 1 using trigonometry. Let $\beta=\frac{1}{2} \angle A B C$ and $\gamma=\frac{1}{2} \angle A C B$. Observe that $\angle F I B=\angle A I B-90^{\circ}=\gamma$. Hence, $\frac{B F}{F I}=\frac{\sin \angle F I B}{\sin \angle I B F}=\frac{\sin \gamma}{\sin \beta}$. Similarly, $\frac{C E}{E I}=\frac{\sin \beta}{\sin \gamma}$. As $F I=E I$, we get

$$
\begin{equation*}
\frac{B F}{C E}=\frac{B F}{F I} \cdot \frac{E I}{C E}=\left(\frac{\sin \gamma}{\sin \beta}\right)^{2} \tag{5}
\end{equation*}
$$

Together with (1) and (2), we find that

$$
\frac{B D^{\prime}}{C D^{\prime}}=\frac{A C}{A B} \cdot\left(\frac{\sin \gamma}{\sin \beta}\right)^{2}=\frac{\sin 2 \beta}{\sin 2 \gamma} \cdot\left(\frac{\sin \gamma}{\sin \beta}\right)^{2}=\frac{\tan \gamma}{\tan \beta}=\frac{I D / C D}{I D / B D}=\frac{B D}{C D}
$$

This shows $D^{\prime}=D$ and the result follows.
Solution 2. Let $\omega_{A}$ be the $A$-mixtilinear incircle of triangle $A B C$. From the properties of mixtilinear incircles, $\omega_{A}$ touches sides $A B$ and $A C$ at $F$ and $E$ respectively. Suppose $\omega_{A}$ is tangent to $\Gamma$ at $T$. Let $A M$ meet $\Gamma$ again at $Y$, and let $D_{1}, T_{1}$ be the reflections of $D$ and $T$ with respect to the perpendicular bisector of $B C$ respectively. It is well-known that $\angle B A T=\angle D_{1} A C$ so that $A, D_{1}, T_{1}$ are collinear.


We then show that $X, M, T_{1}$ are collinear. Let $R$ be the radical centre of $\omega_{A}, \Gamma$ and the circumcircle of triangle $A E F$. Then $R$ lies on $A X, E F$ and the tangent at $T$ to $\Gamma$. Let $A T$ meet $\omega_{A}$ again at $S$ and meet $E F$ at $P$. Obviously, $S F T E$ is a harmonic quadrilateral. Projecting from $T$, the pencil $(R, P ; F, E)$ is harmonic. We further project the pencil onto $\Gamma$ from $A$, so that $X B T C$ is a harmonic quadrilateral. As $T T_{1} / / B C$, the projection from $T_{1}$ onto $B C$ maps $T$ to a point at infinity, and hence maps $X$ to the midpoint of $B C$, which is $M$. This shows $X, M, T_{1}$ are collinear.

We have two ways to finish the proof.

- Method 1. Note that both $A Y$ and $X T_{1}$ are chords of $\Gamma$ passing through the midpoint $M$ of the chord $B C$. By the Butterfly Theorem, $X Y$ and $A T_{1}$ cut $B C$ at a pair of symmetric points with respect to $M$, and hence $X, D, Y$ are collinear. The proof is thus complete.
- Method 2. Here, we finish the proof without using the Butterfly Theorem. As $D T T_{1} D_{1}$ is an isosceles trapezoid, we have

$$
\measuredangle Y T D=\measuredangle Y T T_{1}+\measuredangle T_{1} T D=\measuredangle Y A T_{1}+\measuredangle A D_{1} D=\measuredangle Y M D
$$

so that $D, T, Y, M$ are concyclic. As $X, M, T_{1}$ are collinear, we have

$$
\measuredangle A Y D=\measuredangle M T D=\measuredangle D_{1} T_{1} M=\measuredangle A T_{1} X=\measuredangle A Y X
$$

This shows $X, D, Y$ are collinear.

G3. Let $B=(-1,0)$ and $C=(1,0)$ be fixed points on the coordinate plane. A nonempty, bounded subset $S$ of the plane is said to be nice if
(i) there is a point $T$ in $S$ such that for every point $Q$ in $S$, the segment $T Q$ lies entirely in $S$; and
(ii) for any triangle $P_{1} P_{2} P_{3}$, there exists a unique point $A$ in $S$ and a permutation $\sigma$ of the indices $\{1,2,3\}$ for which triangles $A B C$ and $P_{\sigma(1)} P_{\sigma(2)} P_{\sigma(3)}$ are similar.

Prove that there exist two distinct nice subsets $S$ and $S^{\prime}$ of the set $\{(x, y): x \geqslant 0, y \geqslant 0\}$ such that if $A \in S$ and $A^{\prime} \in S^{\prime}$ are the unique choices of points in (ii), then the product $B A \cdot B A^{\prime}$ is a constant independent of the triangle $P_{1} P_{2} P_{3}$.

Solution. If in the similarity of $\triangle A B C$ and $\triangle P_{\sigma(1)} P_{\sigma(2)} P_{\sigma(3)}, B C$ corresponds to the longest side of $\triangle P_{1} P_{2} P_{3}$, then we have $B C \geqslant A B \geqslant A C$. The condition $B C \geqslant A B$ is equivalent to $(x+1)^{2}+y^{2} \leqslant 4$, while $A B \geqslant A C$ is trivially satisfied for any point in the first quadrant. Then we first define

$$
S=\left\{(x, y):(x+1)^{2}+y^{2} \leqslant 4, x \geqslant 0, y \geqslant 0\right\} .
$$

Note that $S$ is the intersection of a disk and the first quadrant, so it is bounded and convex, and we can choose any $T \in S$ to meet condition (i). For any point $A$ in $S$, the relation $B C \geqslant A B \geqslant A C$ always holds. Therefore, the point $A$ in (ii) is uniquely determined, while its existence is guaranteed by the above construction.


Next, if in the similarity of $\triangle A^{\prime} B C$ and $\triangle P_{\sigma(1)} P_{\sigma(2)} P_{\sigma(3)}, B C$ corresponds to the second longest side of $\triangle P_{1} P_{2} P_{3}$, then we have $A^{\prime} B \geqslant B C \geqslant A^{\prime} C$. The two inequalities are equivalent to $(x+1)^{2}+y^{2} \geqslant 4$ and $(x-1)^{2}+y^{2} \leqslant 4$ respectively. Then we define

$$
S^{\prime}=\left\{(x, y):(x+1)^{2}+y^{2} \geqslant 4,(x-1)^{2}+y^{2} \leqslant 4, x \geqslant 0, y \geqslant 0\right\} .
$$

The boundedness condition is satisfied while (ii) can be argued as in the previous case. For (i), note that $S^{\prime}$ contains points inside the disk $(x-1)^{2}+y^{2} \leqslant 4$ and outside the disk $(x+1)^{2}+y^{2} \geqslant 4$. This shows we can take $T^{\prime}=(1,2)$ in (i), which is the topmost point of the circle $(x-1)^{2}+y^{2}=4$.

It remains to check that the product $B A \cdot B A^{\prime}$ is a constant. Suppose we are given a triangle $P_{1} P_{2} P_{3}$ with $P_{1} P_{2} \geqslant P_{2} P_{3} \geqslant P_{3} P_{1}$. By the similarity, we have

$$
B A=B C \cdot \frac{P_{2} P_{3}}{P_{1} P_{2}} \quad \text { and } \quad B A^{\prime}=B C \cdot \frac{P_{1} P_{2}}{P_{2} P_{3}}
$$

Thus $B A \cdot B A^{\prime}=B C^{2}=4$, which is certainly independent of the triangle $P_{1} P_{2} P_{3}$.
Comment. The original version of this problem includes the condition that the interiors of $S$ and $S^{\prime}$ are disjoint. We remove this condition since it is hard to define the interior of a point set rigorously.

G4. Let $A B C$ be a triangle with $A B=A C \neq B C$ and let $I$ be its incentre. The line $B I$ meets $A C$ at $D$, and the line through $D$ perpendicular to $A C$ meets $A I$ at $E$. Prove that the reflection of $I$ in $A C$ lies on the circumcircle of triangle $B D E$.

## Solution 1.



Let $\Gamma$ be the circle with centre $E$ passing through $B$ and $C$. Since $E D \perp A C$, the point $F$ symmetric to $C$ with respect to $D$ lies on $\Gamma$. From $\angle D C I=\angle I C B=\angle C B I$, the line $D C$ is a tangent to the circumcircle of triangle $I B C$. Let $J$ be the symmetric point of $I$ with respect to $D$. Using directed lengths, from

$$
D C \cdot D F=-D C^{2}=-D I \cdot D B=D J \cdot D B
$$

the point $J$ also lies on $\Gamma$. Let $I^{\prime}$ be the reflection of $I$ in $A C$. Since $I J$ and $C F$ bisect each other, $C J F I$ is a parallelogram. From $\angle F I^{\prime} C=\angle C I F=\angle F J C$, we find that $I^{\prime}$ lies on $\Gamma$. This gives $E I^{\prime}=E B$.

Note that $A C$ is the internal angle bisector of $\angle B D I^{\prime}$. This shows $D E$ is the external angle bisector of $\angle B D I^{\prime}$ as $D E \perp A C$. Together with $E I^{\prime}=E B$, it is well-known that $E$ lies on the circumcircle of triangle $B D I^{\prime}$.

Solution 2. Let $I^{\prime}$ be the reflection of $I$ in $A C$ and let $S$ be the intersection of $I^{\prime} C$ and $A I$. Using directed angles, we let $\theta=\measuredangle A C I=\measuredangle I C B=\measuredangle C B I$. We have

$$
\measuredangle I^{\prime} S E=\measuredangle I^{\prime} C A+\measuredangle C A I=\theta+\left(\frac{\pi}{2}+2 \theta\right)=3 \theta+\frac{\pi}{2}
$$

and

$$
\measuredangle I^{\prime} D E=\measuredangle I^{\prime} D C+\frac{\pi}{2}=\measuredangle C D I+\frac{\pi}{2}=\measuredangle D C B+\measuredangle C B D+\frac{\pi}{2}=3 \theta+\frac{\pi}{2}
$$

This shows $I^{\prime}, D, E, S$ are concyclic.
Next, we find $\measuredangle I^{\prime} S B=2 \measuredangle I^{\prime} S E=6 \theta$ and $\measuredangle I^{\prime} D B=2 \measuredangle C D I=6 \theta$. Therefore, $I^{\prime}, D, B, S$ are concyclic so that $I^{\prime}, D, E, B, S$ lie on the same circle. The result then follows.


Comment. The point $S$ constructed in Solution 2 may lie on the same side as $A$ of $B C$. Also, since $S$ lies on the circumcircle of the non-degenerate triangle $B D E$, we implicitly know that $S$ is not an ideal point. Indeed, one can verify that $I^{\prime} C$ and $A I$ are parallel if and only if triangle $A B C$ is equilateral.
Solution 3. Let $I^{\prime}$ be the reflection of $I$ in $A C$, and let $D^{\prime}$ be the second intersection of $A I$ and the circumcircle of triangle $A B D$. Since $A D^{\prime}$ bisects $\angle B A D$, point $D^{\prime}$ is the midpoint of the arc $B D$ and $D D^{\prime}=B D^{\prime}=C D^{\prime}$. Obviously, $A, E, D^{\prime}$ lie on $A I$ in this order.


We find that $\angle E D^{\prime} D=\angle A D^{\prime} D=\angle A B D=\angle I B C=\angle I C B$. Next, since $D^{\prime}$ is the circumcentre of triangle $B C D$, we have $\angle E D D^{\prime}=90^{\circ}-\angle D^{\prime} D C=\angle C B D=\angle I B C$. The two relations show that triangles $E D^{\prime} D$ and $I C B$ are similar. Therefore, we have

$$
\frac{B C}{C I^{\prime}}=\frac{B C}{C I}=\frac{D D^{\prime}}{D^{\prime} E}=\frac{B D^{\prime}}{D^{\prime} E}
$$

Also, we get

$$
\angle B C I^{\prime}=\angle B C A+\angle A C I^{\prime}=\angle B C A+\angle I C A=\angle B C A+\angle D B C=\angle B D A=\angle B D^{\prime} E
$$

These show triangles $B C I^{\prime}$ and $B D^{\prime} E$ are similar, and hence triangles $B C D^{\prime}$ and $B I^{\prime} E$ are similar. As $B C D^{\prime}$ is isosceles, we obtain $B E=I^{\prime} E$.

As $D E$ is the external angle bisector of $\angle B D I^{\prime}$ and $E I^{\prime}=E B$, we know that $E$ lies on the circumcircle of triangle $B D I^{\prime}$.

Solution 4. Let $A I$ and $B I$ meet the circumcircle of triangle $A B C$ again at $A^{\prime}$ and $B^{\prime}$ respectively, and let $E^{\prime}$ be the reflection of $E$ in $A C$. From

$$
\begin{aligned}
\angle B^{\prime} A E^{\prime} & =\angle B^{\prime} A D-\angle E^{\prime} A D=\frac{\angle A B C}{2}-\frac{\angle B A C}{2}=90^{\circ}-\angle B A C-\frac{\angle A B C}{2} \\
& =90^{\circ}-\angle B^{\prime} D A=\angle B^{\prime} D E^{\prime},
\end{aligned}
$$

points $B^{\prime}, A, D, E^{\prime}$ are concyclic. Then

$$
\angle D B^{\prime} E^{\prime}=\angle D A E^{\prime}=\frac{\angle B A C}{2}=\angle B A A^{\prime}=\angle D B^{\prime} A^{\prime}
$$

and hence $B^{\prime}, E^{\prime}, A^{\prime}$ are collinear. It is well-known that $A^{\prime} B^{\prime}$ is the perpendicular bisector of $C I$, so that $C E^{\prime}=I E^{\prime}$. Let $I^{\prime}$ be the reflection of $I$ in $A C$. This implies $B E=C E=I^{\prime} E$. As $D E$ is the external angle bisector of $\angle B D I^{\prime}$ and $E I^{\prime}=E B$, we know that $E$ lies on the circumcircle of triangle $B D I^{\prime}$.


Solution 5. Let $F$ be the intersection of $C I$ and $A B$. Clearly, $F$ and $D$ are symmetric with respect to $A I$. Let $O$ be the circumcentre of triangle $B I F$, and let $I^{\prime}$ be the reflection of $I$ in $A C$.


From $\angle B F O=90^{\circ}-\angle F I B=\frac{1}{2} \angle B A C=\angle B A I$, we get $E I / / F O$. Also, from the relation $\angle O I B=90^{\circ}-\angle B F I=90^{\circ}-\angle C D I=\angle I^{\prime} I D$, we know that $O, I, I^{\prime}$ are collinear.

Note that $E D / / O I$ since both are perpendicular to $A C$. Then $\angle F E I=\angle D E I=\angle O I E$. Together with $E I / / F O$, the quadrilateral $E F O I$ is an isosceles trapezoid. Therefore, we find that $\angle D I E=\angle F I E=\angle O E I$ so $O E / / I D$. Then $D E O I$ is a parallelogram. Hence, we have $D I^{\prime}=D I=E O$, which shows $D E O I^{\prime}$ is an isosceles trapezoid. In addition, $E D=O I=O B$ and $O E / / B D$ imply $E O B D$ is another isosceles trapezoid. In particular, both $D E O I^{\prime}$ and $E O B D$ are cyclic. This shows $B, D, E, I^{\prime}$ are concyclic.

Solution 6. Let $I^{\prime}$ be the reflection of $I$ in $A C$. Denote by $T$ and $M$ the projections from $I$ to sides $A B$ and $B C$ respectively. Since $B I$ is the perpendicular bisector of $T M$, we have

$$
\begin{equation*}
D T=D M \tag{1}
\end{equation*}
$$

Since $\angle A D E=\angle A T I=90^{\circ}$ and $\angle D A E=\angle T A I$, we have $\triangle A D E \sim \triangle A T I$. This shows $\frac{A D}{A E}=\frac{A T}{A I}=\frac{A T}{A I^{\prime}}$. Together with $\angle D A T=2 \angle D A E=\angle E A I^{\prime}$, this yields $\triangle D A T \sim \triangle E A I^{\prime}$. In particular, we have

$$
\begin{equation*}
\frac{D T}{E I^{\prime}}=\frac{A T}{A I^{\prime}}=\frac{A T}{A I} \tag{2}
\end{equation*}
$$

Obviously, the right-angled triangles $A M B$ and $A T I$ are similar. Then we get

$$
\begin{equation*}
\frac{A M}{A B}=\frac{A T}{A I} \tag{3}
\end{equation*}
$$

Next, from $\triangle A M B \sim \triangle A T I \sim \triangle A D E$, we have $\frac{A M}{A B}=\frac{A D}{A E}$ so that $\triangle A D M \sim \triangle A E B$. It follows that

$$
\begin{equation*}
\frac{D M}{E B}=\frac{A M}{A B} . \tag{4}
\end{equation*}
$$

Combining (1), (2), (3) and (4), we get $E B=E I^{\prime}$. As $D E$ is the external angle bisector of $\angle B D I^{\prime}$, we know that $E$ lies on the circumcircle of triangle $B D I^{\prime}$.


Comment. A stronger version of this problem is to ask the contestants to prove the reflection of $I$ in $A C$ lies on the circumcircle of triangle $B D E$ if and only if $A B=A C$. Some of the above solutions can be modified to prove the converse statement to the original problem. For example, we borrow some ideas from Solution 2 to establish the converse as follows.


Let $I^{\prime}$ be the reflection of $I$ in $A C$ and suppose $B, E, D, I^{\prime}$ lie on a circle $\Gamma$. Let $A I$ meet $\Gamma$ again at $S$. As $D E$ is the external angle bisector of $\angle B D I^{\prime}$, we have $E B=E I^{\prime}$. Using directed angles, we get

$$
\measuredangle C I^{\prime} S=\measuredangle C I^{\prime} D+\measuredangle D I^{\prime} S=\measuredangle D I C+\measuredangle D E S=\measuredangle D I C+\measuredangle D E A=\measuredangle D I C+\measuredangle D C B=0 .
$$

This means $I^{\prime}, C, S$ are collinear. From this we get $\measuredangle B S E=\measuredangle E S I^{\prime}=\measuredangle E S C$ and hence $A S$ bisects both $\angle B A C$ and $\angle B S C$. Clearly, $S$ and $A$ are distinct points. It follows that $\triangle B A S \cong \triangle C A S$ and thus $A B=A C$.

As in some of the above solutions, an obvious way to prove the stronger version is to establish the following equivalence: $B E=E I^{\prime}$ if and only if $A B=A C$. In addition to the ideas used in those solutions, one can use trigonometry to express the lengths of $B E$ and $E I^{\prime}$ in terms of the side lengths of triangle $A B C$ and to establish the equivalence.

G5. Let $D$ be the foot of perpendicular from $A$ to the Euler line (the line passing through the circumcentre and the orthocentre) of an acute scalene triangle $A B C$. A circle $\omega$ with centre $S$ passes through $A$ and $D$, and it intersects sides $A B$ and $A C$ at $X$ and $Y$ respectively. Let $P$ be the foot of altitude from $A$ to $B C$, and let $M$ be the midpoint of $B C$. Prove that the circumcentre of triangle $X S Y$ is equidistant from $P$ and $M$.

Solution 1. Denote the orthocentre and circumcentre of triangle $A B C$ by $H$ and $O$ respectively. Let $Q$ be the midpoint of $A H$ and $N$ be the nine-point centre of triangle $A B C$. It is known that $Q$ lies on the nine-point circle of triangle $A B C, N$ is the midpoint of $Q M$ and that $Q M$ is parallel to $A O$.

Let the perpendicular from $S$ to $X Y$ meet line $Q M$ at $S^{\prime}$. Let $E$ be the foot of altitude from $B$ to side $A C$. Since $Q$ and $S$ lie on the perpendicular bisector of $A D$, using directed angles, we have

$$
\begin{aligned}
\measuredangle S D Q & =\measuredangle Q A S=\measuredangle X A S-\measuredangle X A Q=\left(\frac{\pi}{2}-\measuredangle A Y X\right)-\measuredangle B A P=\measuredangle C B A-\measuredangle A Y X \\
& =(\measuredangle C B A-\measuredangle A C B)-\measuredangle B C A-\measuredangle A Y X=\measuredangle P E M-(\measuredangle B C A+\measuredangle A Y X) \\
& =\measuredangle P Q M-\measuredangle(B C, X Y)=\frac{\pi}{2}-\measuredangle\left(S^{\prime} Q, B C\right)-\angle(B C, X Y)=\measuredangle S S^{\prime} Q
\end{aligned}
$$

This shows $D, S^{\prime}, S, Q$ are concyclic.


Let the perpendicular from $N$ to $B C$ intersect line $S S^{\prime}$ at $O_{1}$. (Note that the two lines coincide when $S$ is the midpoint of $A O$, in which case the result is true since the circumcentre of triangle $X S Y$ must lie on this line.) It suffices to show that $O_{1}$ is the circumcentre of triangle $X S Y$ since $N$ lies on the perpendicular bisector of $P M$. From

$$
\measuredangle D S^{\prime} O_{1}=\measuredangle D Q S=\measuredangle S Q A=\angle(S Q, Q A)=\angle\left(O D, O_{1} N\right)=\measuredangle D N O_{1}
$$

since $S Q / / O D$ and $Q A / / O_{1} N$, we know that $D, O_{1}, S^{\prime}, N$ are concyclic. Therefore, we get

$$
\measuredangle S D S^{\prime}=\measuredangle S Q S^{\prime}=\angle\left(S Q, Q S^{\prime}\right)=\angle\left(N D, N S^{\prime}\right)=\measuredangle D N S^{\prime}
$$

so that $S D$ is a tangent to the circle through $D, O_{1}, S^{\prime}, N$. Then we have

$$
\begin{equation*}
S S^{\prime} \cdot S O_{1}=S D^{2}=S X^{2} \tag{1}
\end{equation*}
$$

Next, we show that $S$ and $S^{\prime}$ are symmetric with respect to $X Y$. By the Sine Law, we have

$$
\frac{S S^{\prime}}{\sin \angle S Q S^{\prime}}=\frac{S Q}{\sin \angle S S^{\prime} Q}=\frac{S Q}{\sin \angle S D Q}=\frac{S Q}{\sin \angle S A Q}=\frac{S A}{\sin \angle S Q A} .
$$

It follows that

$$
S S^{\prime}=S A \cdot \frac{\sin \angle S Q S^{\prime}}{\sin \angle S Q A}=S A \cdot \frac{\sin \angle H O A}{\sin \angle O H A}=S A \cdot \frac{A H}{A O}=S A \cdot 2 \cos A
$$

which is twice the distance from $S$ to $X Y$. Note that $S$ and $C$ lie on the same side of the perpendicular bisector of $P M$ if and only if $\angle S A C<\angle O A C$ if and only if $\angle Y X A>\angle C B A$. This shows $S$ and $O_{1}$ lie on different sides of $X Y$. As $S^{\prime}$ lies on ray $S O_{1}$, it follows that $S$ and $S^{\prime}$ cannot lie on the same side of $X Y$. Therefore, $S$ and $S^{\prime}$ are symmetric with respect to $X Y$.

Let $d$ be the diameter of the circumcircle of triangle $X S Y$. As $S S^{\prime}$ is twice the distance from $S$ to $X Y$ and $S X=S Y$, we have $S S^{\prime}=2 \frac{S X^{2}}{d}$. It follows from (1) that $d=2 S O_{1}$. As $S O_{1}$ is the perpendicular bisector of $X Y$, point $O_{1}$ is the circumcentre of triangle $X S Y$.

Solution 2. Denote the orthocentre and circumcentre of triangle $A B C$ by $H$ and $O$ respectively. Let $O_{1}$ be the circumcentre of triangle $X S Y$. Consider two other possible positions of $S$. We name them $S^{\prime}$ and $S^{\prime \prime}$ and define the analogous points $X^{\prime}, Y^{\prime}, O_{1}^{\prime}, X^{\prime \prime}, Y^{\prime \prime} O_{1}^{\prime \prime}$ accordingly. Note that $S, S^{\prime}, S^{\prime \prime}$ lie on the perpendicular bisector of $A D$.

As $X X^{\prime}$ and $Y Y^{\prime}$ meet at $A$ and the circumcircles of triangles $A X Y$ and $A X^{\prime} Y^{\prime}$ meet at $D$, there is a spiral similarity with centre $D$ mapping $X Y$ to $X^{\prime} Y^{\prime}$. We find that

$$
\measuredangle S X Y=\frac{\pi}{2}-\measuredangle Y A X=\frac{\pi}{2}-\measuredangle Y^{\prime} A X^{\prime}=\measuredangle S^{\prime} X^{\prime} Y^{\prime}
$$

and similarly $\measuredangle S Y X=\measuredangle S^{\prime} Y^{\prime} X^{\prime}$. This shows triangles $S X Y$ and $S^{\prime} X^{\prime} Y^{\prime}$ are directly similar. Then the spiral similarity with centre $D$ takes points $S, X, Y, O_{1}$ to $S^{\prime}, X^{\prime}, Y^{\prime}, O_{1}^{\prime}$. Similarly, there is a spiral similarity with centre $D$ mapping $S, X, Y, O_{1}$ to $S^{\prime \prime}, X^{\prime \prime}, Y^{\prime \prime}, O_{1}^{\prime \prime}$. From these, we see that there is a spiral similarity taking the corresponding points $S, S^{\prime}, S^{\prime \prime}$ to points $O_{1}, O_{1}^{\prime}, O_{1}^{\prime \prime}$. In particular, $O_{1}, O_{1}^{\prime}, O_{1}^{\prime \prime}$ are collinear.


It now suffices to show that $O_{1}$ lies on the perpendicular bisector of $P M$ for two special cases.

Firstly, we take $S$ to be the midpoint of $A H$. Then $X$ and $Y$ are the feet of altitudes from $C$ and $B$ respectively. It is well-known that the circumcircle of triangle $X S Y$ is the nine-point circle of triangle $A B C$. Then $O_{1}$ is the nine-point centre and $O_{1} P=O_{1} M$. Indeed, $P$ and $M$ also lies on the nine-point circle.

Secondly, we take $S^{\prime}$ to be the midpoint of $A O$. Then $X^{\prime}$ and $Y^{\prime}$ are the midpoints of $A B$ and $A C$ respectively. Then $X^{\prime} Y^{\prime} / / B C$. Clearly, $S^{\prime}$ lies on the perpendicular bisector of $P M$. This shows the perpendicular bisectors of $X^{\prime} Y^{\prime}$ and $P M$ coincide. Hence, we must have $O_{1}^{\prime} P=O_{1}^{\prime} M$.


G6. Let $A B C D$ be a convex quadrilateral with $\angle A B C=\angle A D C<90^{\circ}$. The internal angle bisectors of $\angle A B C$ and $\angle A D C$ meet $A C$ at $E$ and $F$ respectively, and meet each other at point $P$. Let $M$ be the midpoint of $A C$ and let $\omega$ be the circumcircle of triangle $B P D$. Segments $B M$ and $D M$ intersect $\omega$ again at $X$ and $Y$ respectively. Denote by $Q$ the intersection point of lines $X E$ and $Y F$. Prove that $P Q \perp A C$.

## Solution 1.



Let $\omega_{1}$ be the circumcircle of triangle $A B C$. We first prove that $Y$ lies on $\omega_{1}$. Let $Y^{\prime}$ be the point on ray $M D$ such that $M Y^{\prime} \cdot M D=M A^{2}$. Then triangles $M A Y^{\prime}$ and $M D A$ are oppositely similar. Since $M C^{2}=M A^{2}=M Y^{\prime} \cdot M D$, triangles $M C Y^{\prime}$ and $M D C$ are also oppositely similar. Therefore, using directed angles, we have

$$
\measuredangle A Y^{\prime} C=\measuredangle A Y^{\prime} M+\measuredangle M Y^{\prime} C=\measuredangle M A D+\measuredangle D C M=\measuredangle C D A=\measuredangle A B C
$$

so that $Y^{\prime}$ lies on $\omega_{1}$.
Let $Z$ be the intersection point of lines $B C$ and $A D$. Since $\measuredangle P D Z=\measuredangle P B C=\measuredangle P B Z$, point $Z$ lies on $\omega$. In addition, from $\measuredangle Y^{\prime} B Z=\measuredangle Y^{\prime} B C=\measuredangle Y^{\prime} A C=\measuredangle Y^{\prime} A M=\measuredangle Y^{\prime} D Z$, we also know that $Y^{\prime}$ lies on $\omega$. Note that $\angle A D C$ is acute implies $M A \neq M D$ so $M Y^{\prime} \neq M D$. Therefore, $Y^{\prime}$ is the second intersection of $D M$ and $\omega$. Then $Y^{\prime}=Y$ and hence $Y$ lies on $\omega_{1}$.

Next, by the Angle Bisector Theorem and the similar triangles, we have

$$
\frac{F A}{F C}=\frac{A D}{C D}=\frac{A D}{A M} \cdot \frac{C M}{C D}=\frac{Y A}{Y M} \cdot \frac{Y M}{Y C}=\frac{Y A}{Y C} .
$$

Hence, $F Y$ is the internal angle bisector of $\angle A Y C$.
Let $B^{\prime}$ be the second intersection of the internal angle bisector of $\angle C B A$ and $\omega_{1}$. Then $B^{\prime}$ is the midpoint of arc $A C$ not containing $B$. Therefore, $Y B^{\prime}$ is the external angle bisector of $\angle A Y C$, so that $B^{\prime} Y \perp F Y$.

Denote by $l$ the line through $P$ parallel to $A C$. Suppose $l$ meets line $B^{\prime} Y$ at $S$. From

$$
\begin{aligned}
\measuredangle P S Y & =\measuredangle\left(A C, B^{\prime} Y\right)=\measuredangle A C Y+\measuredangle C Y B^{\prime}=\measuredangle A C Y+\measuredangle C A B^{\prime}=\measuredangle A C Y+\measuredangle B^{\prime} C A \\
& =\measuredangle B^{\prime} C Y=\measuredangle B^{\prime} B Y=\measuredangle P B Y
\end{aligned}
$$

the point $S$ lies on $\omega$. Similarly, the line through $X$ perpendicular to $X E$ also passes through the second intersection of $l$ and $\omega$, which is the point $S$. From $Q Y \perp Y S$ and $Q X \perp X S$, point $Q$ lies on $\omega$ and $Q S$ is a diameter of $\omega$. Therefore, $P Q \perp P S$ so that $P Q \perp A C$.

Solution 2. Denote by $\omega_{1}$ and $\omega_{2}$ the circumcircles of triangles $A B C$ and $A D C$ respectively. Since $\angle A B C=\angle A D C$, we know that $\omega_{1}$ and $\omega_{2}$ are symmetric with respect to the midpoint $M$ of $A C$.

Firstly, we show that $X$ lies on $\omega_{2}$. Let $X_{1}$ be the second intersection of ray $M B$ and $\omega_{2}$ and $X^{\prime}$ be its symmetric point with respect to $M$. Then $X^{\prime}$ lies on $\omega_{1}$ and $X^{\prime} A X_{1} C$ is a parallelogram. Hence, we have

$$
\begin{aligned}
\measuredangle D X_{1} B & =\measuredangle D X_{1} A+\measuredangle A X_{1} B=\measuredangle D C A+\measuredangle A X_{1} X^{\prime}=\measuredangle D C A+\measuredangle C X^{\prime} X_{1} \\
& =\measuredangle D C A+\measuredangle C A B=\measuredangle(C D, A B) .
\end{aligned}
$$



Also, we have

$$
\measuredangle D P B=\measuredangle P D C+\angle(C D, A B)+\measuredangle A B P=\angle(C D, A B)
$$

These yield $\measuredangle D X_{1} B=\measuredangle D P B$ and hence $X_{1}$ lies on $\omega$. It follows that $X_{1}=X$ and $X$ lies on $\omega_{2}$. Similarly, $Y$ lies on $\omega_{1}$.

Next, we prove that $Q$ lies on $\omega$. Suppose the perpendicular bisector of $A C$ meet $\omega_{1}$ at $B^{\prime}$ and $M_{1}$ and meet $\omega_{2}$ at $D^{\prime}$ and $M_{2}$, so that $B, M_{1}$ and $D^{\prime}$ lie on the same side of $A C$. Note that $B^{\prime}$ lies on the angle bisector of $\angle A B C$ and similarly $D^{\prime}$ lies on $D P$.

If we denote the area of $W_{1} W_{2} W_{3}$ by $\left[W_{1} W_{2} W_{3}\right.$ ], then

$$
\frac{B A \cdot X^{\prime} A}{B C \cdot X^{\prime} C}=\frac{\frac{1}{2} B A \cdot X^{\prime} A \sin \angle B A X^{\prime}}{\frac{1}{2} B C \cdot X^{\prime} C \sin \angle B C X^{\prime}}=\frac{\left[B A X^{\prime}\right]}{\left[B C X^{\prime}\right]}=\frac{M A}{M C}=1 .
$$

As $B E$ is the angle bisector of $\angle A B C$, we have

$$
\frac{E A}{E C}=\frac{B A}{B C}=\frac{X^{\prime} C}{X^{\prime} A}=\frac{X A}{X C} .
$$

Therefore, $X E$ is the angle bisector of $\angle A X C$, so that $M_{2}$ lies on the line joining $X, E, Q$. Analogously, $M_{1}, F, Q, Y$ are collinear. Thus,

$$
\begin{aligned}
\measuredangle X Q Y & =\measuredangle M_{2} Q M_{1}=\measuredangle Q M_{2} M_{1}+\measuredangle M_{2} M_{1} Q=\measuredangle X M_{2} D^{\prime}+\measuredangle B^{\prime} M_{1} Y \\
& =\measuredangle X D D^{\prime}+\measuredangle B^{\prime} B Y=\measuredangle X D P+\measuredangle P B Y=\measuredangle X B P+\measuredangle P B Y=\measuredangle X B Y,
\end{aligned}
$$

which implies $Q$ lies on $\omega$.
Finally, as $M_{1}$ and $M_{2}$ are symmetric with respect to $M$, the quadrilateral $X^{\prime} M_{2} X M_{1}$ is a parallelogram. Consequently,

$$
\measuredangle X Q P=\measuredangle X B P=\measuredangle X^{\prime} B B^{\prime}=\measuredangle X^{\prime} M_{1} B^{\prime}=\measuredangle X M_{2} M_{1} .
$$

This shows $Q P / / M_{2} M_{1}$. As $M_{2} M_{1} \perp A C$, we get $Q P \perp A C$.
Solution 3. We first state two results which will be needed in our proof.

- Claim 1. In $\triangle X^{\prime} Y^{\prime} Z^{\prime}$ with $X^{\prime} Y^{\prime} \neq X^{\prime} Z^{\prime}$, let $N^{\prime}$ be the midpoint of $Y^{\prime} Z^{\prime}$ and $W^{\prime}$ be the foot of internal angle bisector from $X^{\prime}$. Then $\tan ^{2} \measuredangle W^{\prime} X^{\prime} Z^{\prime}=\tan \measuredangle N^{\prime} X^{\prime} W^{\prime} \tan \measuredangle Z^{\prime} W^{\prime} X^{\prime}$.

Proof.


Without loss of generality, assume $X^{\prime} Y^{\prime}>X^{\prime} Z^{\prime}$. Then $W^{\prime}$ lies between $N^{\prime}$ and $Z^{\prime}$. The signs of both sides agree so it suffices to establish the relation for ordinary angles. Let $\angle W^{\prime} X^{\prime} Z^{\prime}=\alpha, \angle N^{\prime} X^{\prime} W^{\prime}=\beta$ and $\angle Z^{\prime} W^{\prime} X^{\prime}=\gamma$. We have

$$
\frac{\sin (\gamma-\alpha)}{\sin (\alpha-\beta)}=\frac{N^{\prime} X^{\prime}}{N^{\prime} Y^{\prime}}=\frac{N^{\prime} X^{\prime}}{N^{\prime} Z^{\prime}}=\frac{\sin (\gamma+\alpha)}{\sin (\alpha+\beta)}
$$

This implies

$$
\frac{\tan \gamma-\tan \alpha}{\tan \gamma+\tan \alpha}=\frac{\sin \gamma \cos \alpha-\cos \gamma \sin \alpha}{\sin \gamma \cos \alpha+\cos \gamma \sin \alpha}=\frac{\sin \alpha \cos \beta-\cos \alpha \sin \beta}{\sin \alpha \cos \beta+\cos \alpha \sin \beta}=\frac{\tan \alpha-\tan \beta}{\tan \alpha+\tan \beta}
$$

Expanding and simplifying, we get the desired result $\tan ^{2} \alpha=\tan \beta \tan \gamma$.

- Claim 2. Let $A^{\prime} B^{\prime} C^{\prime} D^{\prime}$ be a quadrilateral inscribed in circle $\Gamma$. Let diagonals $A^{\prime} C^{\prime}$ and $B^{\prime} D^{\prime}$ meet at $E^{\prime}$, and $F^{\prime}$ be the intersection of lines $A^{\prime} B^{\prime}$ and $C^{\prime} D^{\prime}$. Let $M^{\prime}$ be the midpoint of $E^{\prime} F^{\prime}$. Then the power of $M^{\prime}$ with respect to $\Gamma$ is equal to $\left(M^{\prime} E^{\prime}\right)^{2}$.

Proof.


Let $O^{\prime}$ be the centre of $\Gamma$ and let $\Gamma^{\prime}$ be the circle with centre $M^{\prime}$ passing through $E^{\prime}$. Let $F_{1}$ be the inversion image of $F^{\prime}$ with respect to $\Gamma$. It is well-known that $E^{\prime}$ lies on the polar of $F^{\prime}$ with respect to $\Gamma$. This shows $E^{\prime} F_{1} \perp O^{\prime} F^{\prime}$ and hence $F_{1}$ lies on $\Gamma^{\prime}$. It follows that the inversion image of $\Gamma^{\prime}$ with respect to $\Gamma$ is $\Gamma^{\prime}$ itself. This shows $\Gamma^{\prime}$ is orthogonal to $\Gamma$, and thus the power of $M^{\prime}$ with respect to $\Gamma$ is the square of radius of $\Gamma^{\prime}$, which is $\left(M^{\prime} E^{\prime}\right)^{2}$.

We return to the main problem. Let $Z$ be the intersection of lines $A D$ and $B C$, and $W$ be the intersection of lines $A B$ and $C D$. Since $\measuredangle P D Z=\measuredangle P B C=\measuredangle P B Z$, point $Z$ lies on $\omega$. Similarly, $W$ lies on $\omega$. Applying Claim 2 to the cyclic quadrilateral $Z B D W$, we know that the power of $M$ with respect to $\omega$ is $M A^{2}$. Hence, $M X \cdot M B=M A^{2}$.

Suppose the line through $B$ perpendicular to $B E$ meets line $A C$ at $T$. Then $B E$ and $B T$ are the angle bisectors of $\angle C B A$. This shows $(T, E ; A, C)$ is harmonic. Thus, we have $M E \cdot M T=M A^{2}=M X \cdot M B$. It follows that $E, T, B, X$ are concyclic.


The result is trivial for the special case $A D=C D$ since $P, Q$ lie on the perpendicular bisector of $A C$ in that case. Similarly, the case $A B=C B$ is trivial. It remains to consider the general cases where we can apply Claim 1 in the latter part of the proof.

Let the projections from $P$ and $Q$ to $A C$ be $P^{\prime}$ and $Q^{\prime}$ respectively. Then $P Q \perp A C$ if and only if $P^{\prime}=Q^{\prime}$ if and only if $\frac{E P^{\prime}}{F P^{\prime}}=\frac{E Q^{\prime}}{F Q^{\prime}}$ in terms of directed lengths. Note that

$$
\frac{E P^{\prime}}{F P^{\prime}}=\frac{\tan \measuredangle E F P}{\tan \measuredangle F E P}=\frac{\tan \measuredangle A F D}{\tan \measuredangle A E B} .
$$

Next, we have $\frac{E Q^{\prime}}{F Q^{\prime}}=\frac{\tan \measuredangle E F Q}{\tan \angle F E Q}$ where $\measuredangle F E Q=\measuredangle T E X=\measuredangle T B X=\frac{\pi}{2}+\measuredangle E B M$ and by symmetry $\measuredangle E F Q=\frac{\pi}{2}+\measuredangle F D M$. Combining all these, it suffices to show

$$
\frac{\tan \measuredangle A F D}{\tan \measuredangle A E B}=\frac{\tan \measuredangle M B E}{\tan \measuredangle M D F} .
$$

We now apply Claim 1 twice to get

$$
\tan \measuredangle A F D \tan \measuredangle M D F=\tan ^{2} \measuredangle F D C=\tan ^{2} \measuredangle E B A=\tan \measuredangle M B E \tan \measuredangle A E B .
$$

The result then follows.

G7. Let $I$ be the incentre of a non-equilateral triangle $A B C, I_{A}$ be the $A$-excentre, $I_{A}^{\prime}$ be the reflection of $I_{A}$ in $B C$, and $l_{A}$ be the reflection of line $A I_{A}^{\prime}$ in $A I$. Define points $I_{B}, I_{B}^{\prime}$ and line $l_{B}$ analogously. Let $P$ be the intersection point of $l_{A}$ and $l_{B}$.
(a) Prove that $P$ lies on line $O I$ where $O$ is the circumcentre of triangle $A B C$.
(b) Let one of the tangents from $P$ to the incircle of triangle $A B C$ meet the circumcircle at points $X$ and $Y$. Show that $\angle X I Y=120^{\circ}$.

## Solution 1.

(a) Let $A^{\prime}$ be the reflection of $A$ in $B C$ and let $M$ be the second intersection of line $A I$ and the circumcircle $\Gamma$ of triangle $A B C$. As triangles $A B A^{\prime}$ and $A O C$ are isosceles with $\angle A B A^{\prime}=2 \angle A B C=\angle A O C$, they are similar to each other. Also, triangles $A B I_{A}$ and $A I C$ are similar. Therefore we have

$$
\frac{A A^{\prime}}{A I_{A}}=\frac{A A^{\prime}}{A B} \cdot \frac{A B}{A I_{A}}=\frac{A C}{A O} \cdot \frac{A I}{A C}=\frac{A I}{A O}
$$

Together with $\angle A^{\prime} A I_{A}=\angle I A O$, we find that triangles $A A^{\prime} I_{A}$ and $A I O$ are similar.


Denote by $P^{\prime}$ the intersection of line $A P$ and line $O I$. Using directed angles, we have

$$
\begin{aligned}
\measuredangle M A P^{\prime} & =\measuredangle I_{A}^{\prime} A I_{A}=\measuredangle I_{A}^{\prime} A A^{\prime}-\measuredangle I_{A} A A^{\prime}=\measuredangle A A^{\prime} I_{A}-\measuredangle(A M, O M) \\
& =\measuredangle A I O-\measuredangle A M O=\measuredangle M O P^{\prime} .
\end{aligned}
$$

This shows $M, O, A, P^{\prime}$ are concyclic.

Denote by $R$ and $r$ the circumradius and inradius of triangle $A B C$. Then

$$
I P^{\prime}=\frac{I A \cdot I M}{I O}=\frac{I O^{2}-R^{2}}{I O}
$$

is independent of $A$. Hence, $B P$ also meets line $O I$ at the same point $P^{\prime}$ so that $P^{\prime}=P$, and $P$ lies on $O I$.
(b) By Poncelet's Porism, the other tangents to the incircle of triangle $A B C$ from $X$ and $Y$ meet at a point $Z$ on $\Gamma$. Let $T$ be the touching point of the incircle to $X Y$, and let $D$ be the midpoint of $X Y$. We have

$$
\begin{aligned}
O D & =I T \cdot \frac{O P}{I P}=r\left(1+\frac{O I}{I P}\right)=r\left(1+\frac{O I^{2}}{O I \cdot I P}\right)=r\left(1+\frac{R^{2}-2 R r}{R^{2}-I O^{2}}\right) \\
& =r\left(1+\frac{R^{2}-2 R r}{2 R r}\right)=\frac{R}{2}=\frac{O X}{2} .
\end{aligned}
$$

This shows $\angle X Z Y=60^{\circ}$ and hence $\angle X I Y=120^{\circ}$.

## Solution 2.

(a) Note that triangles $A I_{B} C$ and $I_{A} B C$ are similar since their corresponding interior angles are equal. Therefore, the four triangles $A I_{B}^{\prime} C, A I_{B} C, I_{A} B C$ and $I_{A}^{\prime} B C$ are all similar. From $\triangle A I_{B}^{\prime} C \sim \triangle I_{A}^{\prime} B C$, we get $\triangle A I_{A}^{\prime} C \sim \triangle I_{B}^{\prime} B C$. From $\measuredangle A B P=\measuredangle I_{B}^{\prime} B C=\measuredangle A I_{A}^{\prime} C$ and $\measuredangle B A P=\measuredangle I_{A}^{\prime} A C$, the triangles $A B P$ and $A I_{A}^{\prime} C$ are directly similar.


Consider the inversion with centre $A$ and radius $\sqrt{A B \cdot A C}$ followed by the reflection in $A I$. Then $B$ and $C$ are mapped to each other, and $I$ and $I_{A}$ are mapped to each other.

From the similar triangles obtained, we have $A P \cdot A I_{A}^{\prime}=A B \cdot A C$ so that $P$ is mapped to $I_{A}^{\prime}$ under the transformation. In addition, line $A O$ is mapped to the altitude from $A$, and hence $O$ is mapped to the reflection of $A$ in $B C$, which we call point $A^{\prime}$. Note that $A A^{\prime} I_{A} I_{A}^{\prime}$ is an isosceles trapezoid, which shows it is inscribed in a circle. The preimage of this circle is a straight line, meaning that $O, I, P$ are collinear.
(b) Denote by $R$ and $r$ the circumradius and inradius of triangle $A B C$. Note that by the above transformation, we have $\triangle A P O \sim \triangle A A^{\prime} I_{A}^{\prime}$ and $\triangle A A^{\prime} I_{A} \sim \triangle A I O$. Therefore, we find that

$$
P O=A^{\prime} I_{A}^{\prime} \cdot \frac{A O}{A I_{A}^{\prime}}=A I_{A} \cdot \frac{A O}{A^{\prime} I_{A}}=\frac{A I_{A}}{A^{\prime} I_{A}} \cdot A O=\frac{A O}{I O} \cdot A O .
$$

This shows $P O \cdot I O=R^{2}$, and it follows that $P$ and $I$ are mapped to each other under the inversion with respect to the circumcircle $\Gamma$ of triangle $A B C$. Then $P X \cdot P Y$, which is the power of $P$ with respect to $\Gamma$, equals $P I \cdot P O$. This yields $X, I, O, Y$ are concyclic.

Let $T$ be the touching point of the incircle to $X Y$, and let $D$ be the midpoint of $X Y$. Then

$$
O D=I T \cdot \frac{P O}{P I}=r \cdot \frac{P O}{P O-I O}=r \cdot \frac{R^{2}}{R^{2}-I O^{2}}=r \cdot \frac{R^{2}}{2 R r}=\frac{R}{2} .
$$

This shows $\angle D O X=60^{\circ}$ and hence $\angle X I Y=\angle X O Y=120^{\circ}$.
Comment. A simplification of this problem is to ask part (a) only. Note that the question in part (b) implicitly requires $P$ to lie on $O I$, or otherwise the angle is not uniquely determined as we can find another tangent from $P$ to the incircle.

G8. Let $A_{1}, B_{1}$ and $C_{1}$ be points on sides $B C, C A$ and $A B$ of an acute triangle $A B C$ respectively, such that $A A_{1}, B B_{1}$ and $C C_{1}$ are the internal angle bisectors of triangle $A B C$. Let $I$ be the incentre of triangle $A B C$, and $H$ be the orthocentre of triangle $A_{1} B_{1} C_{1}$. Show that

$$
A H+B H+C H \geqslant A I+B I+C I
$$

Solution. Without loss of generality, assume $\alpha=\angle B A C \leqslant \beta=\angle C B A \leqslant \gamma=\angle A C B$. Denote by $a, b, c$ the lengths of $B C, C A, A B$ respectively. We first show that triangle $A_{1} B_{1} C_{1}$ is acute.

Choose points $D$ and $E$ on side $B C$ such that $B_{1} D / / A B$ and $B_{1} E$ is the internal angle bisector of $\angle B B_{1} C$. As $\angle B_{1} D B=180^{\circ}-\beta$ is obtuse, we have $B B_{1}>B_{1} D$. Thus,

$$
\frac{B E}{E C}=\frac{B B_{1}}{B_{1} C}>\frac{D B_{1}}{B_{1} C}=\frac{B A}{A C}=\frac{B A_{1}}{A_{1} C} .
$$

Therefore, $B E>B A_{1}$ and $\frac{1}{2} \angle B B_{1} C=\angle B B_{1} E>\angle B B_{1} A_{1}$. Similarly, $\frac{1}{2} \angle B B_{1} A>\angle B B_{1} C_{1}$. It follows that

$$
\angle A_{1} B_{1} C_{1}=\angle B B_{1} A_{1}+\angle B B_{1} C_{1}<\frac{1}{2}\left(\angle B B_{1} C+\angle B B_{1} A\right)=90^{\circ}
$$

is acute. By symmetry, triangle $A_{1} B_{1} C_{1}$ is acute.
Let $B B_{1}$ meet $A_{1} C_{1}$ at $F$. From $\alpha \leqslant \gamma$, we get $a \leqslant c$, which implies

$$
B A_{1}=\frac{c a}{b+c} \leqslant \frac{a c}{a+b}=B C_{1}
$$

and hence $\angle B C_{1} A_{1} \leqslant \angle B A_{1} C_{1}$. As $B F$ is the internal angle bisector of $\angle A_{1} B C_{1}$, this shows $\angle B_{1} F C_{1}=\angle B F A_{1} \leqslant 90^{\circ}$. Hence, $H$ lies on the same side of $B B_{1}$ as $C_{1}$. This shows $H$ lies inside triangle $B B_{1} C_{1}$. Similarly, from $\alpha \leqslant \beta$ and $\beta \leqslant \gamma$, we know that $H$ lies inside triangles $C C_{1} B_{1}$ and $A A_{1} C_{1}$.


As $\alpha \leqslant \beta \leqslant \gamma$, we have $\alpha \leqslant 60^{\circ} \leqslant \gamma$. Then $\angle B I C \leqslant 120^{\circ} \leqslant \angle A I B$. Firstly, suppose $\angle A I C \geqslant 120^{\circ}$.

Rotate points $B, I, H$ through $60^{\circ}$ about $A$ to $B^{\prime}, I^{\prime}, H^{\prime}$ so that $B^{\prime}$ and $C$ lie on different sides of $A B$. Since triangle $A I^{\prime} I$ is equilateral, we have

$$
\begin{equation*}
A I+B I+C I=I^{\prime} I+B^{\prime} I^{\prime}+I C=B^{\prime} I^{\prime}+I^{\prime} I+I C . \tag{1}
\end{equation*}
$$

Similarly,

$$
\begin{equation*}
A H+B H+C H=H^{\prime} H+B^{\prime} H^{\prime}+H C=B^{\prime} H^{\prime}+H^{\prime} H+H C \tag{2}
\end{equation*}
$$

As $\angle A I I^{\prime}=\angle A I^{\prime} I=60^{\circ}, \angle A I^{\prime} B^{\prime}=\angle A I B \geqslant 120^{\circ}$ and $\angle A I C \geqslant 120^{\circ}$, the quadrilateral $B^{\prime} I^{\prime} I C$ is convex and lies on the same side of $B^{\prime} C$ as $A$.

Next, since $H$ lies inside triangle $A C C_{1}, H$ lies outside $B^{\prime} I^{\prime} I C$. Also, $H$ lying inside triangle $A B I$ implies $H^{\prime}$ lies inside triangle $A B^{\prime} I^{\prime}$. This shows $H^{\prime}$ lies outside $B^{\prime} I^{\prime} I C$ and hence the convex quadrilateral $B^{\prime} I^{\prime} I C$ is contained inside the quadrilateral $B^{\prime} H^{\prime} H C$. It follows that the perimeter of $B^{\prime} I^{\prime} I C$ cannot exceed the perimeter of $B^{\prime} H^{\prime} H C$. From (1) and (2), we conclude that

$$
A H+B H+C H \geqslant A I+B I+C I
$$

For the case $\angle A I C<120^{\circ}$, we can rotate $B, I, H$ through $60^{\circ}$ about $C$ to $B^{\prime}, I^{\prime}, H^{\prime}$ so that $B^{\prime}$ and $A$ lie on different sides of $B C$. The proof is analogous to the previous case and we still get the desired inequality.

## Number Theory

N1. For any positive integer $k$, denote the sum of digits of $k$ in its decimal representation by $S(k)$. Find all polynomials $P(x)$ with integer coefficients such that for any positive integer $n \geqslant 2016$, the integer $P(n)$ is positive and

$$
\begin{equation*}
S(P(n))=P(S(n)) \tag{1}
\end{equation*}
$$

## Answer.

- $P(x)=c$ where $1 \leqslant c \leqslant 9$ is an integer; or
- $P(x)=x$.

Solution 1. We consider three cases according to the degree of $P$.

- Case 1. $P(x)$ is a constant polynomial.

Let $P(x)=c$ where $c$ is an integer constant. Then (1) becomes $S(c)=c$. This holds if and only if $1 \leqslant c \leqslant 9$.

- Case 2. $\operatorname{deg} P=1$.

We have the following observation. For any positive integers $m, n$, we have

$$
\begin{equation*}
S(m+n) \leqslant S(m)+S(n) \tag{2}
\end{equation*}
$$

and equality holds if and only if there is no carry in the addition $m+n$.
Let $P(x)=a x+b$ for some integers $a, b$ where $a \neq 0$. As $P(n)$ is positive for large $n$, we must have $a \geqslant 1$. The condition (1) becomes $S(a n+b)=a S(n)+b$ for all $n \geqslant 2016$. Setting $n=2025$ and $n=2020$ respectively, we get

$$
S(2025 a+b)-S(2020 a+b)=(a S(2025)+b)-(a S(2020)+b)=5 a
$$

On the other hand, (2) implies

$$
S(2025 a+b)=S((2020 a+b)+5 a) \leqslant S(2020 a+b)+S(5 a)
$$

These give $5 a \leqslant S(5 a)$. As $a \geqslant 1$, this holds only when $a=1$, in which case (1) reduces to $S(n+b)=S(n)+b$ for all $n \geqslant 2016$. Then we find that

$$
\begin{equation*}
S(n+1+b)-S(n+b)=(S(n+1)+b)-(S(n)+b)=S(n+1)-S(n) . \tag{3}
\end{equation*}
$$

If $b>0$, we choose $n$ such that $n+1+b=10^{k}$ for some sufficiently large $k$. Note that all the digits of $n+b$ are 9 's, so that the left-hand side of (3) equals $1-9 k$. As $n$ is a positive integer less than $10^{k}-1$, we have $S(n)<9 k$. Therefore, the right-hand side of (3) is at least $1-(9 k-1)=2-9 k$, which is a contradiction.

The case $b<0$ can be handled similarly by considering $n+1$ to be a large power of 10 . Therefore, we conclude that $P(x)=x$, in which case (1) is trivially satisfied.

- Case 3. $\operatorname{deg} P \geqslant 2$.

Suppose the leading term of $P$ is $a_{d} n^{d}$ where $a_{d} \neq 0$. Clearly, we have $a_{d}>0$. Consider $n=10^{k}-1$ in (1). We get $S(P(n))=P(9 k)$. Note that $P(n)$ grows asymptotically as fast as $n^{d}$, so $S(P(n))$ grows asymptotically as no faster than a constant multiple of $k$. On the other hand, $P(9 k)$ grows asymptotically as fast as $k^{d}$. This shows the two sides of the last equation cannot be equal for sufficiently large $k$ since $d \geqslant 2$.

Therefore, we conclude that $P(x)=c$ where $1 \leqslant c \leqslant 9$ is an integer, or $P(x)=x$.
Solution 2. Let $P(x)=a_{d} x^{d}+a_{d-1} x^{d-1}+\cdots+a_{0}$. Clearly $a_{d}>0$. There exists an integer $m \geqslant 1$ such that $\left|a_{i}\right|<10^{m}$ for all $0 \leqslant i \leqslant d$. Consider $n=9 \times 10^{k}$ for a sufficiently large integer $k$ in (1). If there exists an index $0 \leqslant i \leqslant d-1$ such that $a_{i}<0$, then all digits of $P(n)$ in positions from $10^{i k+m+1}$ to $10^{(i+1) k-1}$ are all 9's. Hence, we have $S(P(n)) \geqslant 9(k-m-1)$. On the other hand, $P(S(n))=P(9)$ is a fixed constant. Therefore, (1) cannot hold for large $k$. This shows $a_{i} \geqslant 0$ for all $0 \leqslant i \leqslant d-1$.

Hence, $P(n)$ is an integer formed by the nonnegative integers $a_{d} \times 9^{d}, a_{d-1} \times 9^{d-1}, \ldots, a_{0}$ by inserting some zeros in between. This yields

$$
S(P(n))=S\left(a_{d} \times 9^{d}\right)+S\left(a_{d-1} \times 9^{d-1}\right)+\cdots+S\left(a_{0}\right)
$$

Combining with (1), we have

$$
S\left(a_{d} \times 9^{d}\right)+S\left(a_{d-1} \times 9^{d-1}\right)+\cdots+S\left(a_{0}\right)=P(9)=a_{d} \times 9^{d}+a_{d-1} \times 9^{d-1}+\cdots+a_{0}
$$

As $S(m) \leqslant m$ for any positive integer $m$, with equality when $1 \leqslant m \leqslant 9$, this forces each $a_{i} \times 9^{i}$ to be a positive integer between 1 and 9 . In particular, this shows $a_{i}=0$ for $i \geqslant 2$ and hence $d \leqslant 1$. Also, we have $a_{1} \leqslant 1$ and $a_{0} \leqslant 9$. If $a_{1}=1$ and $1 \leqslant a_{0} \leqslant 9$, we take $n=10^{k}+\left(10-a_{0}\right)$ for sufficiently large $k$ in (1). This yields a contradiction since

$$
S(P(n))=S\left(10^{k}+10\right)=2 \neq 11=P\left(11-a_{0}\right)=P(S(n))
$$

The zero polynomial is also rejected since $P(n)$ is positive for large $n$. The remaining candidates are $P(x)=x$ or $P(x)=a_{0}$ where $1 \leqslant a_{0} \leqslant 9$, all of which satisfy (1), and hence are the only solutions.

N2. Let $\tau(n)$ be the number of positive divisors of $n$. Let $\tau_{1}(n)$ be the number of positive divisors of $n$ which have remainders 1 when divided by 3 . Find all possible integral values of the fraction $\frac{\tau(10 n)}{\tau_{1}(10 n)}$.

Answer. All composite numbers together with 2.
Solution. In this solution, we always use $p_{i}$ to denote primes congruent to $1 \bmod 3$, and use $q_{j}$ to denote primes congruent to $2 \bmod 3$. When we express a positive integer $m$ using its prime factorization, we also include the special case $m=1$ by allowing the exponents to be zeros. We first compute $\tau_{1}(m)$ for a positive integer $m$.

- Claim. Let $m=3^{x} p_{1}^{a_{1}} p_{2}^{a_{2}} \cdots p_{s}^{a_{s}} q_{1}^{b_{1}} q_{2}^{b_{2}} \cdots q_{t}^{b_{t}}$ be the prime factorization of $m$. Then

$$
\begin{equation*}
\tau_{1}(m)=\prod_{i=1}^{s}\left(a_{i}+1\right)\left\lceil\frac{1}{2} \prod_{j=1}^{t}\left(b_{j}+1\right)\right\rceil . \tag{1}
\end{equation*}
$$

Proof. To choose a divisor of $m$ congruent to $1 \bmod 3$, it cannot have the prime divisor 3, while there is no restriction on choosing prime factors congruent to $1 \bmod 3$. Also, we have to choose an even number of prime factors (counted with multiplicity) congruent to $2 \bmod 3$.

If $\prod_{j=1}^{t}\left(b_{j}+1\right)$ is even, then we may assume without loss of generality $b_{1}+1$ is even. We can choose the prime factors $q_{2}, q_{3}, \ldots, q_{t}$ freely in $\prod_{j=2}^{t}\left(b_{j}+1\right)$ ways. Then the parity of the number of $q_{1}$ is uniquely determined, and hence there are $\frac{1}{2}\left(b_{1}+1\right)$ ways to choose the exponent of $q_{1}$. Hence (1) is verified in this case.

If $\prod_{j=1}^{t}\left(b_{j}+1\right)$ is odd, we use induction on $t$ to count the number of choices. When $t=1$, there are $\left\lceil\frac{b_{1}+1}{2}\right\rceil$ choices for which the exponent is even and $\left\lfloor\frac{b_{1}+1}{2}\right\rfloor$ choices for which the exponent is odd. For the inductive step, we find that there are

$$
\left\lceil\frac{1}{2} \prod_{j=1}^{t-1}\left(b_{j}+1\right)\right\rceil \cdot\left\lceil\frac{b_{t}+1}{2}\right\rceil+\left\lfloor\frac{1}{2} \prod_{j=1}^{t-1}\left(b_{j}+1\right)\right\rfloor \cdot\left\lfloor\frac{b_{t}+1}{2}\right\rfloor=\left\lceil\frac{1}{2} \prod_{j=1}^{t}\left(b_{j}+1\right)\right\rceil
$$

choices with an even number of prime factors and hence $\left\lfloor\frac{1}{2} \prod_{j=1}^{t}\left(b_{j}+1\right)\right\rfloor$ choices with an odd number of prime factors. Hence (1) is also true in this case.

Let $n=3^{x} 2^{y} 5^{z} p_{1}^{a_{1}} p_{2}^{a_{2}} \cdots p_{s}^{a_{s}} q_{1}^{b_{1}} q_{2}^{b_{2}} \cdots q_{t}^{b_{t}}$. Using the well-known formula for computing the divisor function, we get

$$
\begin{equation*}
\tau(10 n)=(x+1)(y+2)(z+2) \prod_{i=1}^{s}\left(a_{i}+1\right) \prod_{j=1}^{t}\left(b_{j}+1\right) . \tag{2}
\end{equation*}
$$

By the Claim, we have

$$
\begin{equation*}
\tau_{1}(10 n)=\prod_{i=1}^{s}\left(a_{i}+1\right)\left\lceil\frac{1}{2}(y+2)(z+2) \prod_{j=1}^{t}\left(b_{j}+1\right)\right\rceil . \tag{3}
\end{equation*}
$$

If $c=(y+2)(z+2) \prod_{j=1}^{t}\left(b_{j}+1\right)$ is even, then (2) and (3) imply

$$
\frac{\tau(10 n)}{\tau_{1}(10 n)}=2(x+1)
$$

In this case $\frac{\tau(10 n)}{\tau_{1}(10 n)}$ can be any even positive integer as $x$ runs through all nonnegative integers.
If $c$ is odd, which means $y, z$ are odd and each $b_{j}$ is even, then (2) and (3) imply

$$
\begin{equation*}
\frac{\tau(10 n)}{\tau_{1}(10 n)}=\frac{2(x+1) c}{c+1} \tag{4}
\end{equation*}
$$

For this to be an integer, we need $c+1$ divides $2(x+1)$ since $c$ and $c+1$ are relatively prime. Let $2(x+1)=k(c+1)$. Then (4) reduces to

$$
\begin{equation*}
\frac{\tau(10 n)}{\tau_{1}(10 n)}=k c=k(y+2)(z+2) \prod_{j=1}^{t}\left(b_{j}+1\right) . \tag{5}
\end{equation*}
$$

Noting that $y, z$ are odd, the integers $y+2$ and $z+2$ are at least 3. This shows the integer in this case must be composite. On the other hand, for any odd composite number $a b$ with $a, b \geqslant 3$, we may simply take $n=3^{\frac{a b-1}{2}} \cdot 2^{a-2} \cdot 5^{b-2}$ so that $\frac{\tau(10 n)}{\tau_{1}(10 n)}=a b$ from (5).

We conclude that the fraction can be any even integer or any odd composite number. Equivalently, it can be 2 or any composite number.

N3. Define $P(n)=n^{2}+n+1$. For any positive integers $a$ and $b$, the set

$$
\{P(a), P(a+1), P(a+2), \ldots, P(a+b)\}
$$

is said to be fragrant if none of its elements is relatively prime to the product of the other elements. Determine the smallest size of a fragrant set.

Answer. 6.
Solution. We have the following observations.
(i) $(P(n), P(n+1))=1$ for any $n$.

We have $(P(n), P(n+1))=\left(n^{2}+n+1, n^{2}+3 n+3\right)=\left(n^{2}+n+1,2 n+2\right)$. Noting that $n^{2}+n+1$ is odd and $\left(n^{2}+n+1, n+1\right)=(1, n+1)=1$, the claim follows.
(ii) $(P(n), P(n+2))=1$ for $n \not \equiv 2(\bmod 7)$ and $(P(n), P(n+2))=7$ for $n \equiv 2(\bmod 7)$.

From $(2 n+7) P(n)-(2 n-1) P(n+2)=14$ and the fact that $P(n)$ is odd, $(P(n), P(n+2))$ must be a divisor of 7 . The claim follows by checking $n \equiv 0,1, \ldots, 6(\bmod 7)$ directly.
(iii) $(P(n), P(n+3))=1$ for $n \not \equiv 1(\bmod 3)$ and $3 \mid(P(n), P(n+3))$ for $n \equiv 1(\bmod 3)$.

From $(n+5) P(n)-(n-1) P(n+3)=18$ and the fact that $P(n)$ is odd, $(P(n), P(n+3))$ must be a divisor of 9 . The claim follows by checking $n \equiv 0,1,2(\bmod 3)$ directly.

Suppose there exists a fragrant set with at most 5 elements. We may assume it contains exactly 5 elements $P(a), P(a+1), \ldots, P(a+4)$ since the following argument also works with fewer elements. Consider $P(a+2)$. From (i), it is relatively prime to $P(a+1)$ and $P(a+3)$. Without loss of generality, assume $(P(a), P(a+2))>1$. From (ii), we have $a \equiv 2(\bmod 7)$. The same observation implies $(P(a+1), P(a+3))=1$. In order that the set is fragrant, $(P(a), P(a+3))$ and $(P(a+1), P(a+4))$ must both be greater than 1 . From (iii), this holds only when both $a$ and $a+1$ are congruent to $1 \bmod 3$, which is a contradiction.

It now suffices to construct a fragrant set of size 6. By the Chinese Remainder Theorem, we can take a positive integer $a$ such that

$$
a \equiv 7 \quad(\bmod 19), \quad a+1 \equiv 2 \quad(\bmod 7), \quad a+2 \equiv 1 \quad(\bmod 3) .
$$

For example, we may take $a=197$. From (ii), both $P(a+1)$ and $P(a+3)$ are divisible by 7. From (iii), both $P(a+2)$ and $P(a+5)$ are divisible by 3 . One also checks from $19 \mid P(7)=57$ and $19 \mid P(11)=133$ that $P(a)$ and $P(a+4)$ are divisible by 19 . Therefore, the set $\{P(a), P(a+1), \ldots, P(a+5)\}$ is fragrant.

Therefore, the smallest size of a fragrant set is 6 .
Comment. "Fragrant Harbour" is the English translation of "Hong Kong".
A stronger version of this problem is to show that there exists a fragrant set of size $k$ for any $k \geqslant 6$. We present a proof here.

For each even positive integer $m$ which is not divisible by 3 , since $m^{2}+3 \equiv 3(\bmod 4)$, we can find a prime $p_{m} \equiv 3(\bmod 4)$ such that $p_{m} \mid m^{2}+3$. Clearly, $p_{m}>3$.

If $b=2 t \geqslant 6$, we choose $a$ such that $3 \mid 2(a+t)+1$ and $p_{m} \mid 2(a+t)+1$ for each $1 \leqslant m \leqslant b$ with $m \equiv 2,4(\bmod 6)$. For $0 \leqslant r \leqslant t$ and $3 \mid r$, we have $a+t \pm r \equiv 1(\bmod 3)$ so that $3 \mid P(a+t \pm r)$. For $0 \leqslant r \leqslant t$ and $(r, 3)=1$, we have

$$
4 P(a+t \pm r) \equiv(-1 \pm 2 r)^{2}+2(-1 \pm 2 r)+4=4 r^{2}+3 \equiv 0 \quad\left(\bmod p_{2 r}\right)
$$

Hence, $\{P(a), P(a+1), \ldots, P(a+b)\}$ is fragrant.
If $b=2 t+1 \geqslant 7$ (the case $b=5$ has been done in the original problem), we choose $a$ such that $3 \mid 2(a+t)+1$ and $p_{m} \mid 2(a+t)+1$ for $1 \leqslant m \leqslant b$ with $m \equiv 2,4(\bmod 6)$, and that $a+b \equiv 9(\bmod 13)$. Note that $a$ exists by the Chinese Remainder Theorem since $p_{m} \neq 13$ for all $m$. The even case shows that $\{P(a), P(a+1), \ldots, P(a+b-1)\}$ is fragrant. Also, one checks from $13 \mid P(9)=91$ and $13 \mid P(3)=13$ that $P(a+b)$ and $P(a+b-6)$ are divisible by 13. The proof is thus complete.

N4. Let $n, m, k$ and $l$ be positive integers with $n \neq 1$ such that $n^{k}+m n^{l}+1$ divides $n^{k+l}-1$. Prove that

- $m=1$ and $l=2 k$; or
- $l \mid k$ and $m=\frac{n^{k-l}-1}{n^{l}-1}$.

Solution 1. It is given that

$$
\begin{equation*}
n^{k}+m n^{l}+1 \mid n^{k+l}-1 . \tag{1}
\end{equation*}
$$

This implies

$$
\begin{equation*}
n^{k}+m n^{l}+1 \mid\left(n^{k+l}-1\right)+\left(n^{k}+m n^{l}+1\right)=n^{k+l}+n^{k}+m n^{l} . \tag{2}
\end{equation*}
$$

We have two cases to discuss.

- Case 1. $l \geqslant k$.

Since $\left(n^{k}+m n^{l}+1, n\right)=1$, (2) yields

$$
n^{k}+m n^{l}+1 \mid n^{l}+m n^{l-k}+1 .
$$

In particular, we get $n^{k}+m n^{l}+1 \leqslant n^{l}+m n^{l-k}+1$. As $n \geqslant 2$ and $k \geqslant 1,(m-1) n^{l}$ is at least $2(m-1) n^{l-k}$. It follows that the inequality cannot hold when $m \geqslant 2$. For $m=1$, the above divisibility becomes

$$
n^{k}+n^{l}+1 \mid n^{l}+n^{l-k}+1 .
$$

Note that $n^{l}+n^{l-k}+1<n^{l}+n^{l}+1<2\left(n^{k}+n^{l}+1\right)$. Thus we must have $n^{l}+n^{l-k}+1=n^{k}+n^{l}+1$ so that $l=2 k$, which gives the first result.

- Case 2. $l<k$.

This time (2) yields

$$
n^{k}+m n^{l}+1 \mid n^{k}+n^{k-l}+m .
$$

In particular, we get $n^{k}+m n^{l}+1 \leqslant n^{k}+n^{k-l}+m$, which implies

$$
\begin{equation*}
m \leqslant \frac{n^{k-l}-1}{n^{l}-1} \tag{3}
\end{equation*}
$$

On the other hand, from (1) we may let $n^{k+l}-1=\left(n^{k}+m n^{l}+1\right) t$ for some positive integer $t$. Obviously, $t$ is less than $n^{l}$, which means $t \leqslant n^{l}-1$ as it is an integer. Then we have $n^{k+l}-1 \leqslant\left(n^{k}+m n^{l}+1\right)\left(n^{l}-1\right)$, which is the same as

$$
\begin{equation*}
m \geqslant \frac{n^{k-l}-1}{n^{l}-1} \tag{4}
\end{equation*}
$$

Equations (3) and (4) combine to give $m=\frac{n^{k-l}-1}{n^{l}-1}$. As this is an integer, we have $l \mid k-l$. This means $l \mid k$ and it corresponds to the second result.

Solution 2. As in Solution 1, we begin with equation (2).

- Case 1. $l \geqslant k$.

Then (2) yields

$$
n^{k}+m n^{l}+1 \mid n^{l}+m n^{l-k}+1 .
$$

Since $2\left(n^{k}+m n^{l}+1\right)>2 m n^{l}+1>n^{l}+m n^{l-k}+1$, it follows that $n^{k}+m n^{l}+1=n^{l}+m n^{l-k}+1$, that is,

$$
m\left(n^{l}-n^{l-k}\right)=n^{l}-n^{k} .
$$

If $m \geqslant 2$, then $m\left(n^{l}-n^{l-k}\right) \geqslant 2 n^{l}-2 n^{l-k} \geqslant 2 n^{l}-n^{l}>n^{l}-n^{k}$ gives a contradiction. Hence $m=1$ and $l-k=k$, which means $m=1$ and $l=2 k$.

- Case 2. $l<k$.

Then (2) yields

$$
n^{k}+m n^{l}+1 \mid n^{k}+n^{k-l}+m .
$$

Since $2\left(n^{k}+m n^{l}+1\right)>2 n^{k}+m>n^{k}+n^{k-l}+m$, it follows that $n^{k}+m n^{l}+1=n^{k}+n^{k-l}+m$. This gives $m=\frac{n^{k-l}-1}{n^{l}-1}$. Note that $n^{l}-1 \mid n^{k-l}-1$ implies $l \mid k-l$ and hence $l \mid k$. The proof is thus complete.

Comment. Another version of this problem is as follows: let $n, m, k$ and $l$ be positive integers with $n \neq 1$ such that $k$ and $l$ do not divide each other. Show that $n^{k}+m n^{l}+1$ does not divide $n^{k+l}-1$.

N5. Let $a$ be a positive integer which is not a square number. Denote by $A$ the set of all positive integers $k$ such that

$$
\begin{equation*}
k=\frac{x^{2}-a}{x^{2}-y^{2}} \tag{1}
\end{equation*}
$$

for some integers $x$ and $y$ with $x>\sqrt{a}$. Denote by $B$ the set of all positive integers $k$ such that (1) is satisfied for some integers $x$ and $y$ with $0 \leqslant x<\sqrt{a}$. Prove that $A=B$.

Solution 1. We first prove the following preliminary result.

- Claim. For fixed $k$, let $x, y$ be integers satisfying (1). Then the numbers $x_{1}, y_{1}$ defined by

$$
x_{1}=\frac{1}{2}\left(x-y+\frac{(x-y)^{2}-4 a}{x+y}\right), \quad y_{1}=\frac{1}{2}\left(x-y-\frac{(x-y)^{2}-4 a}{x+y}\right)
$$

are integers and satisfy (1) (with $x, y$ replaced by $x_{1}, y_{1}$ respectively).
Proof. Since $x_{1}+y_{1}=x-y$ and

$$
x_{1}=\frac{x^{2}-x y-2 a}{x+y}=-x+\frac{2\left(x^{2}-a\right)}{x+y}=-x+2 k(x-y),
$$

both $x_{1}$ and $y_{1}$ are integers. Let $u=x+y$ and $v=x-y$. The relation (1) can be rewritten as

$$
u^{2}-(4 k-2) u v+\left(v^{2}-4 a\right)=0
$$

By Vieta's Theorem, the number $z=\frac{v^{2}-4 a}{u}$ satisfies

$$
v^{2}-(4 k-2) v z+\left(z^{2}-4 a\right)=0
$$

Since $x_{1}$ and $y_{1}$ are defined so that $v=x_{1}+y_{1}$ and $z=x_{1}-y_{1}$, we can reverse the process and verify (1) for $x_{1}, y_{1}$.

We first show that $B \subset A$. Take any $k \in B$ so that (1) is satisfied for some integers $x, y$ with $0 \leqslant x<\sqrt{a}$. Clearly, $y \neq 0$ and we may assume $y$ is positive. Since $a$ is not a square, we have $k>1$. Hence, we get $0 \leqslant x<y<\sqrt{a}$. Define

$$
x_{1}=\frac{1}{2}\left|x-y+\frac{(x-y)^{2}-4 a}{x+y}\right|, \quad y_{1}=\frac{1}{2}\left(x-y-\frac{(x-y)^{2}-4 a}{x+y}\right) .
$$

By the Claim, $x_{1}$, $y_{1}$ are integers satisfying (1). Also, we have

$$
x_{1} \geqslant-\frac{1}{2}\left(x-y+\frac{(x-y)^{2}-4 a}{x+y}\right)=\frac{2 a+x(y-x)}{x+y} \geqslant \frac{2 a}{x+y}>\sqrt{a} .
$$

This implies $k \in A$ and hence $B \subset A$.

Next, we shall show that $A \subset B$. Take any $k \in A$ so that (1) is satisfied for some integers $x, y$ with $x>\sqrt{a}$. Again, we may assume $y$ is positive. Among all such representations of $k$, we choose the one with smallest $x+y$. Define

$$
x_{1}=\frac{1}{2}\left|x-y+\frac{(x-y)^{2}-4 a}{x+y}\right|, \quad y_{1}=\frac{1}{2}\left(x-y-\frac{(x-y)^{2}-4 a}{x+y}\right) .
$$

By the Claim, $x_{1}, y_{1}$ are integers satisfying (1). Since $k>1$, we get $x>y>\sqrt{a}$. Therefore, we have $y_{1}>\frac{4 a}{x+y}>0$ and $\frac{4 a}{x+y}<x+y$. It follows that

$$
x_{1}+y_{1} \leqslant \max \left\{x-y, \frac{4 a-(x-y)^{2}}{x+y}\right\}<x+y
$$

If $x_{1}>\sqrt{a}$, we get a contradiction due to the minimality of $x+y$. Therefore, we must have $0 \leqslant x_{1}<\sqrt{a}$, which means $k \in B$ so that $A \subset B$.

The two subset relations combine to give $A=B$.
Solution 2. The relation (1) is equivalent to

$$
\begin{equation*}
k y^{2}-(k-1) x^{2}=a \tag{2}
\end{equation*}
$$

Motivated by Pell's Equation, we prove the following, which is essentially the same as the Claim in Solution 1.

- Claim. If $\left(x_{0}, y_{0}\right)$ is a solution to $(2)$, then $\left((2 k-1) x_{0} \pm 2 k y_{0},(2 k-1) y_{0} \pm 2(k-1) x_{0}\right)$ is also a solution to (2).

Proof. We check directly that

$$
\begin{aligned}
& k\left((2 k-1) y_{0} \pm 2(k-1) x_{0}\right)^{2}-(k-1)\left((2 k-1) x_{0} \pm 2 k y_{0}\right)^{2} \\
= & \left(k(2 k-1)^{2}-(k-1)(2 k)^{2}\right) y_{0}^{2}+\left(k(2(k-1))^{2}-(k-1)(2 k-1)^{2}\right) x_{0}^{2} \\
= & k y_{0}^{2}-(k-1) x_{0}^{2}=a
\end{aligned}
$$

If (2) is satisfied for some $0 \leqslant x<\sqrt{a}$ and nonnegative integer $y$, then clearly (1) implies $y>x$. Also, we have $k>1$ since $a$ is not a square number. By the Claim, consider another solution to (2) defined by

$$
x_{1}=(2 k-1) x+2 k y, \quad y_{1}=(2 k-1) y+2(k-1) x .
$$

It satisfies $x_{1} \geqslant(2 k-1) x+2 k(x+1)=(4 k-1) x+2 k>x$. Then we can replace the old solution by a new one which has a larger value in $x$. After a finite number of replacements, we must get a solution with $x>\sqrt{a}$. This shows $B \subset A$.

If (2) is satisfied for some $x>\sqrt{a}$ and nonnegative integer $y$, by the Claim we consider another solution to (2) defined by

$$
x_{1}=|(2 k-1) x-2 k y|, \quad y_{1}=(2 k-1) y-2(k-1) x .
$$

From (2), we get $\sqrt{k} y>\sqrt{k-1} x$. This implies $k y>\sqrt{k(k-1)} x>(k-1) x$ and hence $(2 k-1) x-2 k y<x$. On the other hand, the relation (1) implies $x>y$. Then it is clear that $(2 k-1) x-2 k y>-x$. These combine to give $x_{1}<x$, which means we have found a solution to (2) with $x$ having a smaller absolute value. After a finite number of steps, we shall obtain a solution with $0 \leqslant x<\sqrt{a}$. This shows $A \subset B$.

The desired result follows from $B \subset A$ and $A \subset B$.
Solution 3. It suffices to show $A \cup B$ is a subset of $A \cap B$. We take any $k \in A \cup B$, which means there exist integers $x, y$ satisfying (1). Since $a$ is not a square, it follows that $k \neq 1$. As in Solution 2, the result follows readily once we have proved the existence of a solution $\left(x_{1}, y_{1}\right)$ to (1) with $\left|x_{1}\right|>|x|$, and, in case of $x>\sqrt{a}$, another solution $\left(x_{2}, y_{2}\right)$ with $\left|x_{2}\right|<|x|$.

Without loss of generality, assume $x, y \geqslant 0$. Let $u=x+y$ and $v=x-y$. Then $u \geqslant v$ and (1) becomes

$$
\begin{equation*}
k=\frac{(u+v)^{2}-4 a}{4 u v} . \tag{3}
\end{equation*}
$$

This is the same as

$$
v^{2}+(2 u-4 k u) v+u^{2}-4 a=0 .
$$

Let $v_{1}=4 k u-2 u-v$. Then $u+v_{1}=4 k u-u-v \geqslant 8 u-u-v>u+v$. By Vieta's Theorem, $v_{1}$ satisfies

$$
v_{1}^{2}+(2 u-4 k u) v_{1}+u^{2}-4 a=0
$$

This gives $k=\frac{\left(u+v_{1}\right)^{2}-4 a}{4 u v_{1}}$. As $k$ is an integer, $u+v_{1}$ must be even. Therefore, $x_{1}=\frac{u+v_{1}}{2}$ and $y_{1}=\frac{v_{1}-u}{2}$ are integers. By reversing the process, we can see that $\left(x_{1}, y_{1}\right)$ is a solution to (1), with $x_{1}=\frac{u+v_{1}}{2}>\frac{u+v}{2}=x \geqslant 0$. This completes the first half of the proof.

Suppose $x>\sqrt{a}$. Then $u+v>2 \sqrt{a}$ and (3) can be rewritten as

$$
u^{2}+(2 v-4 k v) u+v^{2}-4 a=0 .
$$

Let $u_{2}=4 k v-2 v-u$. By Vieta's Theorem, we have $u u_{2}=v^{2}-4 a$ and

$$
\begin{equation*}
u_{2}^{2}+(2 v-4 k v) u_{2}+v^{2}-4 a=0 . \tag{4}
\end{equation*}
$$

By $u>0, u+v>2 \sqrt{a}$ and (3), we have $v>0$. If $u_{2} \geqslant 0$, then $v u_{2} \leqslant u u_{2}=v^{2}-4 a<v^{2}$. This shows $u_{2}<v \leqslant u$ and $0<u_{2}+v<u+v$. If $u_{2}<0$, then $\left(u_{2}+v\right)+(u+v)=4 k v>0$ and $u_{2}+v<u+v$ imply $\left|u_{2}+v\right|<u+v$. In any case, since $u_{2}+v$ is even from (4), we can define $x_{2}=\frac{u_{2}+v}{2}$ and $y_{2}=\frac{u_{2}-v}{2}$ so that (1) is satisfied with $\left|x_{2}\right|<x$, as desired. The proof is thus complete.

N6. Denote by $\mathbb{N}$ the set of all positive integers. Find all functions $f: \mathbb{N} \rightarrow \mathbb{N}$ such that for all positive integers $m$ and $n$, the integer $f(m)+f(n)-m n$ is nonzero and divides $m f(m)+n f(n)$.
Answer. $f(n)=n^{2}$ for any $n \in \mathbb{N}$.
Solution. It is given that

$$
\begin{equation*}
f(m)+f(n)-m n \mid m f(m)+n f(n) \tag{1}
\end{equation*}
$$

Taking $m=n=1$ in (1), we have $2 f(1)-1 \mid 2 f(1)$. Then $2 f(1)-1 \mid 2 f(1)-(2 f(1)-1)=1$ and hence $f(1)=1$.

Let $p \geqslant 7$ be a prime. Taking $m=p$ and $n=1$ in (1), we have $f(p)-p+1 \mid p f(p)+1$ and hence

$$
f(p)-p+1 \mid p f(p)+1-p(f(p)-p+1)=p^{2}-p+1
$$

If $f(p)-p+1=p^{2}-p+1$, then $f(p)=p^{2}$. If $f(p)-p+1 \neq p^{2}-p+1$, as $p^{2}-p+1$ is an odd positive integer, we have $p^{2}-p+1 \geqslant 3(f(p)-p+1)$, that is,

$$
\begin{equation*}
f(p) \leqslant \frac{1}{3}\left(p^{2}+2 p-2\right) \tag{2}
\end{equation*}
$$

Taking $m=n=p$ in (1), we have $2 f(p)-p^{2} \mid 2 p f(p)$. This implies

$$
2 f(p)-p^{2} \mid 2 p f(p)-p\left(2 f(p)-p^{2}\right)=p^{3}
$$

By (2) and $f(p) \geqslant 1$, we get

$$
-p^{2}<2 f(p)-p^{2} \leqslant \frac{2}{3}\left(p^{2}+2 p-2\right)-p^{2}<-p
$$

since $p \geqslant 7$. This contradicts the fact that $2 f(p)-p^{2}$ is a factor of $p^{3}$. Thus we have proved that $f(p)=p^{2}$ for all primes $p \geqslant 7$.

Let $n$ be a fixed positive integer. Choose a sufficiently large prime $p$. Consider $m=p$ in (1). We obtain

$$
f(p)+f(n)-p n \mid p f(p)+n f(n)-n(f(p)+f(n)-p n)=p f(p)-n f(p)+p n^{2} .
$$

As $f(p)=p^{2}$, this implies $p^{2}-p n+f(n) \mid p\left(p^{2}-p n+n^{2}\right)$. As $p$ is sufficiently large and $n$ is fixed, $p$ cannot divide $f(n)$, and so $\left(p, p^{2}-p n+f(n)\right)=1$. It follows that $p^{2}-p n+f(n) \mid p^{2}-p n+n^{2}$ and hence

$$
p^{2}-p n+f(n) \mid\left(p^{2}-p n+n^{2}\right)-\left(p^{2}-p n+f(n)\right)=n^{2}-f(n)
$$

Note that $n^{2}-f(n)$ is fixed while $p^{2}-p n+f(n)$ is chosen to be sufficiently large. Therefore, we must have $n^{2}-f(n)=0$ so that $f(n)=n^{2}$ for any positive integer $n$.

Finally, we check that when $f(n)=n^{2}$ for any positive integer $n$, then

$$
f(m)+f(n)-m n=m^{2}+n^{2}-m n
$$

and

$$
m f(m)+n f(n)=m^{3}+n^{3}=(m+n)\left(m^{2}+n^{2}-m n\right)
$$

The latter expression is divisible by the former for any positive integers $m, n$. This shows $f(n)=n^{2}$ is the only solution.

N7. Let $n$ be an odd positive integer. In the Cartesian plane, a cyclic polygon $P$ with area $S$ is chosen. All its vertices have integral coordinates, and the squares of its side lengths are all divisible by $n$. Prove that $2 S$ is an integer divisible by $n$.
Solution. Let $P=A_{1} A_{2} \ldots A_{k}$ and let $A_{k+i}=A_{i}$ for $i \geqslant 1$. By the Shoelace Formula, the area of any convex polygon with integral coordinates is half an integer. Therefore, $2 S$ is an integer. We shall prove by induction on $k \geqslant 3$ that $2 S$ is divisible by $n$. Clearly, it suffices to consider $n=p^{t}$ where $p$ is an odd prime and $t \geqslant 1$.

For the base case $k=3$, let the side lengths of $P$ be $\sqrt{n a}, \sqrt{n b}, \sqrt{n c}$ where $a, b, c$ are positive integers. By Heron's Formula,

$$
16 S^{2}=n^{2}\left(2 a b+2 b c+2 c a-a^{2}-b^{2}-c^{2}\right)
$$

This shows $16 S^{2}$ is divisible by $n^{2}$. Since $n$ is odd, $2 S$ is divisible by $n$.
Assume $k \geqslant 4$. If the square of length of one of the diagonals is divisible by $n$, then that diagonal divides $P$ into two smaller polygons, to which the induction hypothesis applies. Hence we may assume that none of the squares of diagonal lengths is divisible by $n$. As usual, we denote by $\nu_{p}(r)$ the exponent of $p$ in the prime decomposition of $r$. We claim the following.

- Claim. $\nu_{p}\left(A_{1} A_{m}^{2}\right)>\nu_{p}\left(A_{1} A_{m+1}^{2}\right)$ for $2 \leqslant m \leqslant k-1$.

Proof. The case $m=2$ is obvious since $\nu_{p}\left(A_{1} A_{2}^{2}\right) \geqslant p^{t}>\nu_{p}\left(A_{1} A_{3}^{2}\right)$ by the condition and the above assumption.

Suppose $\nu_{p}\left(A_{1} A_{2}^{2}\right)>\nu_{p}\left(A_{1} A_{3}^{2}\right)>\cdots>\nu_{p}\left(A_{1} A_{m}^{2}\right)$ where $3 \leqslant m \leqslant k-1$. For the induction step, we apply Ptolemy's Theorem to the cyclic quadrilateral $A_{1} A_{m-1} A_{m} A_{m+1}$ to get

$$
A_{1} A_{m+1} \times A_{m-1} A_{m}+A_{1} A_{m-1} \times A_{m} A_{m+1}=A_{1} A_{m} \times A_{m-1} A_{m+1}
$$

which can be rewritten as

$$
\begin{align*}
A_{1} A_{m+1}^{2} \times A_{m-1} A_{m}^{2}= & A_{1} A_{m-1}^{2} \times A_{m} A_{m+1}^{2}+A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2} \\
& -2 A_{1} A_{m-1} \times A_{m} A_{m+1} \times A_{1} A_{m} \times A_{m-1} A_{m+1} \tag{1}
\end{align*}
$$

From this, $2 A_{1} A_{m-1} \times A_{m} A_{m+1} \times A_{1} A_{m} \times A_{m-1} A_{m+1}$ is an integer. We consider the component of $p$ of each term in (1). By the inductive hypothesis, we have $\nu_{p}\left(A_{1} A_{m-1}^{2}\right)>\nu_{p}\left(A_{1} A_{m}^{2}\right)$. Also, we have $\nu_{p}\left(A_{m} A_{m+1}^{2}\right) \geqslant p^{t}>\nu_{p}\left(A_{m-1} A_{m+1}^{2}\right)$. These give

$$
\begin{equation*}
\nu_{p}\left(A_{1} A_{m-1}^{2} \times A_{m} A_{m+1}^{2}\right)>\nu_{p}\left(A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2}\right) . \tag{2}
\end{equation*}
$$

Next, we have $\nu_{p}\left(4 A_{1} A_{m-1}^{2} \times A_{m} A_{m+1}^{2} \times A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2}\right)=\nu_{p}\left(A_{1} A_{m-1}^{2} \times A_{m} A_{m+1}^{2}\right)+$ $\nu_{p}\left(A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2}\right)>2 \nu_{p}\left(A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2}\right)$ from (2). This implies

$$
\begin{equation*}
\nu_{p}\left(2 A_{1} A_{m-1} \times A_{m} A_{m+1} \times A_{1} A_{m} \times A_{m-1} A_{m+1}\right)>\nu_{p}\left(A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2}\right) \tag{3}
\end{equation*}
$$

Combining (1), (2) and (3), we conclude that

$$
\nu_{p}\left(A_{1} A_{m+1}^{2} \times A_{m-1} A_{m}^{2}\right)=\nu_{p}\left(A_{1} A_{m}^{2} \times A_{m-1} A_{m+1}^{2}\right)
$$

By $\nu_{p}\left(A_{m-1} A_{m}^{2}\right) \geqslant p^{t}>\nu_{p}\left(A_{m-1} A_{m+1}^{2}\right)$, we get $\nu_{p}\left(A_{1} A_{m+1}^{2}\right)<\nu_{p}\left(A_{1} A_{m}^{2}\right)$. The Claim follows by induction.

From the Claim, we get a chain of inequalities

$$
p^{t}>\nu_{p}\left(A_{1} A_{3}^{2}\right)>\nu_{p}\left(A_{1} A_{4}^{2}\right)>\cdots>\nu_{p}\left(A_{1} A_{k}^{2}\right) \geqslant p^{t}
$$

which yields a contradiction. Therefore, we can show by induction that $2 S$ is divisible by $n$.
Comment. The condition that $P$ is cyclic is crucial. As a counterexample, consider the rhombus with vertices $(0,3),(4,0),(0,-3),(-4,0)$. Each of its squares of side lengths is divisible by 5 , while $2 S=48$ is not.

The proposer also gives a proof for the case $n$ is even. One just needs an extra technical step for the case $p=2$.

N8. Find all polynomials $P(x)$ of odd degree $d$ and with integer coefficients satisfying the following property: for each positive integer $n$, there exist $n$ positive integers $x_{1}, x_{2}, \ldots, x_{n}$ such that $\frac{1}{2}<\frac{P\left(x_{i}\right)}{P\left(x_{j}\right)}<2$ and $\frac{P\left(x_{i}\right)}{P\left(x_{j}\right)}$ is the $d$-th power of a rational number for every pair of indices $i$ and $j$ with $1 \leqslant i, j \leqslant n$.

Answer. $P(x)=a(r x+s)^{d}$ where $a, r, s$ are integers with $a \neq 0, r \geqslant 1$ and $(r, s)=1$.
Solution. Let $P(x)=a_{d} x^{d}+a_{d-1} x^{d-1}+\cdots+a_{0}$. Consider the substitution $y=d a_{d} x+a_{d-1}$. By defining $Q(y)=P(x)$, we find that $Q$ is a polynomial with rational coefficients without the term $y^{d-1}$. Let $Q(y)=b_{d} y^{d}+b_{d-2} y^{d-2}+b_{d-3} y^{d-3}+\cdots+b_{0}$ and $B=\max _{0 \leqslant i \leqslant d}\left\{\left|b_{i}\right|\right\}$ (where $b_{d-1}=0$ ).

The condition shows that for each $n \geqslant 1$, there exist integers $y_{1}, y_{2}, \ldots, y_{n}$ such that $\frac{1}{2}<\frac{Q\left(y_{i}\right)}{Q\left(y_{j}\right)}<2$ and $\frac{Q\left(y_{i}\right)}{Q\left(y_{j}\right)}$ is the $d$-th power of a rational number for $1 \leqslant i, j \leqslant n$. Since $n$ can be arbitrarily large, we may assume all $x_{i}$ 's and hence $y_{i}$ 's are integers larger than some absolute constant in the following.

By Dirichlet's Theorem, since $d$ is odd, we can find a sufficiently large prime $p$ such that $p \equiv 2(\bmod d)$. In particular, we have $(p-1, d)=1$. For this fixed $p$, we choose $n$ to be sufficiently large. Then by the Pigeonhole Principle, there must be $d+1$ of $y_{1}, y_{2}, \ldots, y_{n}$ which are congruent $\bmod p$. Without loss of generality, assume $y_{i} \equiv y_{j}(\bmod p)$ for $1 \leqslant i, j \leqslant d+1$. We shall establish the following.

- Claim. $\frac{Q\left(y_{i}\right)}{Q\left(y_{1}\right)}=\frac{y_{i}^{d}}{y_{1}^{d}}$ for $2 \leqslant i \leqslant d+1$.

Proof. Let $\frac{Q\left(y_{i}\right)}{Q\left(y_{1}\right)}=\frac{l^{d}}{m^{d}}$ where $(l, m)=1$ and $l, m>0$. This can be rewritten in the expanded form

$$
\begin{equation*}
b_{d}\left(m^{d} y_{i}^{d}-l^{d} y_{1}^{d}\right)=-\sum_{j=0}^{d-2} b_{j}\left(m^{d} y_{i}^{j}-l^{d} y_{1}^{j}\right) . \tag{1}
\end{equation*}
$$

Let $c$ be the common denominator of $Q$, so that $c Q(k)$ is an integer for any integer $k$. Note that $c$ depends only on $P$ and so we may assume $(p, c)=1$. Then $y_{1} \equiv y_{i}(\bmod p)$ implies $c Q\left(y_{1}\right) \equiv c Q\left(y_{i}\right)(\bmod p)$.

- Case 1. $p \mid c Q\left(y_{1}\right)$.

In this case, there is a cancellation of $p$ in the numerator and denominator of $\frac{c Q\left(y_{i}\right)}{c Q\left(y_{1}\right)}$, so that $m^{d} \leqslant p^{-1}\left|c Q\left(y_{1}\right)\right|$. Noting $\left|Q\left(y_{1}\right)\right|<2 B y_{1}^{d}$ as $y_{1}$ is large, we get

$$
\begin{equation*}
m \leqslant p^{-\frac{1}{d}}(2 c B)^{\frac{1}{d}} y_{1} \tag{2}
\end{equation*}
$$

For large $y_{1}$ and $y_{i}$, the relation $\frac{1}{2}<\frac{Q\left(y_{i}\right)}{Q\left(y_{1}\right)}<2$ implies

$$
\begin{equation*}
\frac{1}{3}<\frac{y_{i}^{d}}{y_{1}^{d}}<3 \tag{3}
\end{equation*}
$$

We also have

$$
\begin{equation*}
\frac{1}{2}<\frac{l^{d}}{m^{d}}<2 . \tag{4}
\end{equation*}
$$

Now, the left-hand side of (1) is

$$
b_{d}\left(m y_{i}-l y_{1}\right)\left(m^{d-1} y_{i}^{d-1}+m^{d-2} y_{i}^{d-2} l y_{1}+\cdots+l^{d-1} y_{1}^{d-1}\right)
$$

Suppose on the contrary that $m y_{i}-l y_{1} \neq 0$. Then the absolute value of the above expression is at least $\left|b_{d}\right| m^{d-1} y_{i}^{d-1}$. On the other hand, the absolute value of the right-hand side of (1) is at most

$$
\begin{aligned}
\sum_{j=0}^{d-2} B\left(m^{d} y_{i}^{j}+l^{d} y_{1}^{j}\right) & \leqslant(d-1) B\left(m^{d} y_{i}^{d-2}+l^{d} y_{1}^{d-2}\right) \\
& \leqslant(d-1) B\left(7 m^{d} y_{i}^{d-2}\right) \\
& \leqslant 7(d-1) B\left(p^{-\frac{1}{d}}(2 c B)^{\frac{1}{d}} y_{1}\right) m^{d-1} y_{i}^{d-2} \\
& \leqslant 21(d-1) B p^{-\frac{1}{d}}(2 c B)^{\frac{1}{d}} m^{d-1} y_{i}^{d-1}
\end{aligned}
$$

by using successively (3), (4), (2) and again (3). This shows

$$
\left|b_{d}\right| m^{d-1} y_{i}^{d-1} \leqslant 21(d-1) B p^{-\frac{1}{d}}(2 c B)^{\frac{1}{d}} m^{d-1} y_{i}^{d-1},
$$

which is a contradiction for large $p$ as $b_{d}, B, c, d$ depend only on the polynomial $P$. Therefore, we have $m y_{i}-l y_{1}=0$ in this case.

- Case 2. $\left(p, c Q\left(y_{1}\right)\right)=1$.

From $c Q\left(y_{1}\right) \equiv c Q\left(y_{i}\right)(\bmod p)$, we have $l^{d} \equiv m^{d}(\bmod p) . \quad$ Since $(p-1, d)=1$, we use Fermat Little Theorem to conclude $l \equiv m(\bmod p)$. Then $p \mid m y_{i}-l y_{1}$. Suppose on the contrary that $m y_{i}-l y_{1} \neq 0$. Then the left-hand side of (1) has absolute value at least $\left|b_{d}\right| p m^{d-1} y_{i}^{d-1}$. Similar to Case 1, the right-hand side of (1) has absolute value at most

$$
21(d-1) B(2 c B)^{\frac{1}{d}} m^{d-1} y_{i}^{d-1}
$$

which must be smaller than $\left|b_{d}\right| p m^{d-1} y_{i}^{d-1}$ for large $p$. Again this yields a contradiction and hence $m y_{i}-l y_{1}=0$.

In both cases, we find that $\frac{Q\left(y_{i}\right)}{Q\left(y_{1}\right)}=\frac{l^{d}}{m^{d}}=\frac{y_{i}^{d}}{y_{1}^{d}}$.
From the Claim, the polynomial $Q\left(y_{1}\right) y^{d}-y_{1}^{d} Q(y)$ has roots $y=y_{1}, y_{2}, \ldots, y_{d+1}$. Since its degree is at most $d$, this must be the zero polynomial. Hence, $Q(y)=b_{d} y^{d}$. This implies $P(x)=a_{d}\left(x+\frac{a_{d-1}}{d a_{d}}\right)^{d}$. Let $\frac{a_{d-1}}{d a_{d}}=\frac{s}{r}$ with integers $r, s$ where $r \geqslant 1$ and $(r, s)=1$. Since $P$ has integer coefficients, we need $r^{d} \mid a_{d}$. Let $a_{d}=r^{d} a$. Then $P(x)=a(r x+s)^{d}$. It is obvious that such a polynomial satisfies the conditions.

Comment. In the proof, the use of prime and Dirichlet's Theorem can be avoided. One can easily show that each $P\left(x_{i}\right)$ can be expressed in the form $u v_{i}^{d}$ where $u, v_{i}$ are integers and $u$ cannot be divisible by the $d$-th power of a prime (note that $u$ depends only on $P$ ). By fixing a large integer $q$ and by choosing a large $n$, we can apply the Pigeonhole Principle and assume
$x_{1} \equiv x_{2} \equiv \cdots \equiv x_{d+1}(\bmod q)$ and $v_{1} \equiv v_{2} \equiv \cdots \equiv v_{d+1}(\bmod q)$. Then the remaining proof is similar to Case 2 of the Solution.

Alternatively, we give another modification of the proof as follows.
We take a sufficiently large $n$ and consider the corresponding positive integers $y_{1}, y_{2}, \ldots, y_{n}$. For each $2 \leqslant i \leqslant n$, let $\frac{Q\left(y_{i}\right)}{Q\left(y_{1}\right)}=\frac{l_{i}^{d}}{m_{i}^{d}}$.

As in Case 1, if there are $d$ indices $i$ such that the integers $\frac{c\left|Q\left(y_{1}\right)\right|}{m_{i}^{d}}$ are bounded below by a constant depending only on $P$, we can establish the Claim using those $y_{i}$ 's and complete the proof. Similarly, as in Case 2, if there are $d$ indices $i$ such that the integers $\left|m_{i} y_{i}-l_{i} y_{1}\right|$ are bounded below, then the proof goes the same. So it suffices to consider the case where $\frac{c\left|Q\left(y_{1}\right)\right|}{m_{i}^{d}} \leqslant M$ and $\left|m_{i} y_{i}-l_{i} y_{1}\right| \leqslant N$ for all $2 \leqslant i \leqslant n^{\prime}$ where $M, N$ are fixed constants and $n^{\prime}$ is large. Since there are only finitely many choices for $m_{i}$ and $m_{i} y_{i}-l_{i} y_{1}$, by the Pigeonhole Principle, we can assume without loss of generality $m_{i}=m$ and $m_{i} y_{i}-l_{i} y_{1}=t$ for $2 \leqslant i \leqslant d+2$. Then

$$
\frac{Q\left(y_{i}\right)}{Q\left(y_{1}\right)}=\frac{l_{i}^{d}}{m^{d}}=\frac{\left(m y_{i}-t\right)^{d}}{m^{d} y_{1}^{d}}
$$

so that $Q\left(y_{1}\right)(m y-t)^{d}-m^{d} y_{1}^{d} Q(y)$ has roots $y=y_{2}, y_{3}, \ldots, y_{d+2}$. Its degree is at most $d$ and hence it is the zero polynomial. Therefore, $Q(y)=\frac{b_{d}}{m^{d}}(m y-t)^{d}$. Indeed, $Q$ does not have the term $y^{d-1}$, which means $t$ should be 0 . This gives the corresponding $P(x)$ of the desired form.

The two modifications of the Solution work equally well when the degree $d$ is even.

58 ${ }^{\text {th }}$ International Mathematical Olympiad

## Shortlisted Problems (with solutions)

# Shortlisted Problems (with solutions) 

# The Shortlist has to be kept strictly confidential until the conclusion of the following International Mathematical Olympiad. 

IMO General Regulations §6.6

## Contributing Countries

The Organizing Committee and the Problem Selection Committee of IMO 2017 thank the following 51 countries for contributing 150 problem proposals:

Albania, Algeria, Armenia, Australia, Austria, Azerbaijan, Belarus, Belgium, Bulgaria, Cuba, Cyprus, Czech Republic, Denmark, Estonia, France, Georgia, Germany, Greece, Hong Kong, India, Iran, Ireland, Israel, Italy, Japan, Kazakhstan, Latvia, Lithuania, Luxembourg, Mexico, Montenegro, Morocco, Netherlands, Romania, Russia, Serbia, Singapore, Slovakia, Slovenia, South Africa, Sweden, Switzerland, Taiwan, Tajikistan, Tanzania, Thailand, Trinidad and Tobago, Turkey, Ukraine, United Kingdom, U.S.A.

## Problem Selection Committee



Carlos Gustavo Tamm de Araújo Moreira (Gugu) (chairman), Luciano Monteiro de Castro, Ilya I. Bogdanov, Géza Kós, Carlos Yuzo Shine, Zhuo Qun (Alex) Song, Ralph Costa Teixeira, Eduardo Tengan

## Problems

## Algebra

A1. Let $a_{1}, a_{2}, \ldots, a_{n}, k$, and $M$ be positive integers such that

$$
\frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}=k \quad \text { and } \quad a_{1} a_{2} \ldots a_{n}=M
$$

If $M>1$, prove that the polynomial

$$
P(x)=M(x+1)^{k}-\left(x+a_{1}\right)\left(x+a_{2}\right) \cdots\left(x+a_{n}\right)
$$

has no positive roots.
(Trinidad and Tobago)
A2. Let $q$ be a real number. Gugu has a napkin with ten distinct real numbers written on it, and he writes the following three lines of real numbers on the blackboard:

- In the first line, Gugu writes down every number of the form $a-b$, where $a$ and $b$ are two (not necessarily distinct) numbers on his napkin.
- In the second line, Gugu writes down every number of the form $q a b$, where $a$ and $b$ are two (not necessarily distinct) numbers from the first line.
- In the third line, Gugu writes down every number of the form $a^{2}+b^{2}-c^{2}-d^{2}$, where $a, b, c, d$ are four (not necessarily distinct) numbers from the first line.

Determine all values of $q$ such that, regardless of the numbers on Gugu's napkin, every number in the second line is also a number in the third line.
(Austria)
A3. Let $S$ be a finite set, and let $\mathcal{A}$ be the set of all functions from $S$ to $S$. Let $f$ be an element of $\mathcal{A}$, and let $T=f(S)$ be the image of $S$ under $f$. Suppose that $f \circ g \circ f \neq g \circ f \circ g$ for every $g$ in $\mathcal{A}$ with $g \neq f$. Show that $f(T)=T$.
(India)
A4. A sequence of real numbers $a_{1}, a_{2}, \ldots$ satisfies the relation

$$
a_{n}=-\max _{i+j=n}\left(a_{i}+a_{j}\right) \quad \text { for all } n>2017
$$

Prove that this sequence is bounded, i.e., there is a constant $M$ such that $\left|a_{n}\right| \leqslant M$ for all positive integers $n$.

A5. An integer $n \geqslant 3$ is given. We call an $n$-tuple of real numbers $\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ Shiny if for each permutation $y_{1}, y_{2}, \ldots, y_{n}$ of these numbers we have

$$
\sum_{i=1}^{n-1} y_{i} y_{i+1}=y_{1} y_{2}+y_{2} y_{3}+y_{3} y_{4}+\cdots+y_{n-1} y_{n} \geqslant-1
$$

Find the largest constant $K=K(n)$ such that

$$
\sum_{1 \leqslant i<j \leqslant n} x_{i} x_{j} \geqslant K
$$

holds for every Shiny $n$-tuple $\left(x_{1}, x_{2}, \ldots, x_{n}\right)$.
A6. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ such that

$$
f(f(x) f(y))+f(x+y)=f(x y)
$$

for all $x, y \in \mathbb{R}$.
(Albania)
A7. Let $a_{0}, a_{1}, a_{2}, \ldots$ be a sequence of integers and $b_{0}, b_{1}, b_{2}, \ldots$ be a sequence of positive integers such that $a_{0}=0, a_{1}=1$, and

$$
a_{n+1}=\left\{\begin{array}{ll}
a_{n} b_{n}+a_{n-1}, & \text { if } b_{n-1}=1 \\
a_{n} b_{n}-a_{n-1}, & \text { if } b_{n-1}>1
\end{array} \quad \text { for } n=1,2, \ldots\right.
$$

Prove that at least one of the two numbers $a_{2017}$ and $a_{2018}$ must be greater than or equal to 2017 .
(Australia)
A8. Assume that a function $f: \mathbb{R} \rightarrow \mathbb{R}$ satisfies the following condition:
For every $x, y \in \mathbb{R}$ such that $(f(x)+y)(f(y)+x)>0$, we have $f(x)+y=f(y)+x$.
Prove that $f(x)+y \leqslant f(y)+x$ whenever $x>y$.

## Combinatorics

C1. A rectangle $\mathcal{R}$ with odd integer side lengths is divided into small rectangles with integer side lengths. Prove that there is at least one among the small rectangles whose distances from the four sides of $\mathcal{R}$ are either all odd or all even.
(Singapore)
C2. Let $n$ be a positive integer. Define a chameleon to be any sequence of $3 n$ letters, with exactly $n$ occurrences of each of the letters $a, b$, and $c$. Define a swap to be the transposition of two adjacent letters in a chameleon. Prove that for any chameleon $X$, there exists a chameleon $Y$ such that $X$ cannot be changed to $Y$ using fewer than $3 n^{2} / 2$ swaps.
(Australia)
C3. Sir Alex plays the following game on a row of 9 cells. Initially, all cells are empty. In each move, Sir Alex is allowed to perform exactly one of the following two operations:
(1) Choose any number of the form $2^{j}$, where $j$ is a non-negative integer, and put it into an empty cell.
(2) Choose two (not necessarily adjacent) cells with the same number in them; denote that number by $2^{j}$. Replace the number in one of the cells with $2^{j+1}$ and erase the number in the other cell.

At the end of the game, one cell contains the number $2^{n}$, where $n$ is a given positive integer, while the other cells are empty. Determine the maximum number of moves that Sir Alex could have made, in terms of $n$.
(Thailand)
C4. Let $N \geqslant 2$ be an integer. $N(N+1)$ soccer players, no two of the same height, stand in a row in some order. Coach Ralph wants to remove $N(N-1)$ people from this row so that in the remaining row of $2 N$ players, no one stands between the two tallest ones, no one stands between the third and the fourth tallest ones, ..., and finally no one stands between the two shortest ones. Show that this is always possible.
(Russia)
C5. A hunter and an invisible rabbit play a game in the Euclidean plane. The hunter's starting point $H_{0}$ coincides with the rabbit's starting point $R_{0}$. In the $n^{\text {th }}$ round of the game ( $n \geqslant 1$ ), the following happens.
(1) First the invisible rabbit moves secretly and unobserved from its current point $R_{n-1}$ to some new point $R_{n}$ with $R_{n-1} R_{n}=1$.
(2) The hunter has a tracking device (e.g. dog) that returns an approximate position $R_{n}^{\prime}$ of the rabbit, so that $R_{n} R_{n}^{\prime} \leqslant 1$.
(3) The hunter then visibly moves from point $H_{n-1}$ to a new point $H_{n}$ with $H_{n-1} H_{n}=1$.

Is there a strategy for the hunter that guarantees that after $10^{9}$ such rounds the distance between the hunter and the rabbit is below 100 ?

C6. Let $n>1$ be an integer. An $n \times n \times n$ cube is composed of $n^{3}$ unit cubes. Each unit cube is painted with one color. For each $n \times n \times 1$ box consisting of $n^{2}$ unit cubes (of any of the three possible orientations), we consider the set of the colors present in that box (each color is listed only once). This way, we get $3 n$ sets of colors, split into three groups according to the orientation. It happens that for every set in any group, the same set appears in both of the other groups. Determine, in terms of $n$, the maximal possible number of colors that are present.
(Russia)
C7. For any finite sets $X$ and $Y$ of positive integers, denote by $f_{X}(k)$ the $k^{\text {th }}$ smallest positive integer not in $X$, and let

$$
X * Y=X \cup\left\{f_{X}(y): y \in Y\right\}
$$

Let $A$ be a set of $a>0$ positive integers, and let $B$ be a set of $b>0$ positive integers. Prove that if $A * B=B * A$, then

$$
\begin{equation*}
\underbrace{A *(A * \cdots *(A *(A * A)) \cdots)}_{A \text { appears } b \text { times }}=\underbrace{B *(B * \cdots *(B *(B * B)) \cdots)}_{B \text { appears } a \text { times }} . \tag{U.S.A.}
\end{equation*}
$$

C8.
Let $n$ be a given positive integer. In the Cartesian plane, each lattice point with nonnegative coordinates initially contains a butterfly, and there are no other butterflies. The neighborhood of a lattice point $c$ consists of all lattice points within the axis-aligned $(2 n+1) \times(2 n+1)$ square centered at $c$, apart from $c$ itself. We call a butterfly lonely, crowded, or comfortable, depending on whether the number of butterflies in its neighborhood $N$ is respectively less than, greater than, or equal to half of the number of lattice points in $N$.

Every minute, all lonely butterflies fly away simultaneously. This process goes on for as long as there are any lonely butterflies. Assuming that the process eventually stops, determine the number of comfortable butterflies at the final state.

## Geometry

G1. Let $A B C D E$ be a convex pentagon such that $A B=B C=C D, \angle E A B=\angle B C D$, and $\angle E D C=\angle C B A$. Prove that the perpendicular line from $E$ to $B C$ and the line segments $A C$ and $B D$ are concurrent.
(Italy)
G2. Let $R$ and $S$ be distinct points on circle $\Omega$, and let $t$ denote the tangent line to $\Omega$ at $R$. Point $R^{\prime}$ is the reflection of $R$ with respect to $S$. A point $I$ is chosen on the smaller arc $R S$ of $\Omega$ so that the circumcircle $\Gamma$ of triangle $I S R^{\prime}$ intersects $t$ at two different points. Denote by $A$ the common point of $\Gamma$ and $t$ that is closest to $R$. Line $A I$ meets $\Omega$ again at $J$. Show that $J R^{\prime}$ is tangent to $\Gamma$.
(Luxembourg)
G3. Let $O$ be the circumcenter of an acute scalene triangle $A B C$. Line $O A$ intersects the altitudes of $A B C$ through $B$ and $C$ at $P$ and $Q$, respectively. The altitudes meet at $H$. Prove that the circumcenter of triangle $P Q H$ lies on a median of triangle $A B C$.
(Ukraine)
G4. In triangle $A B C$, let $\omega$ be the excircle opposite $A$. Let $D, E$, and $F$ be the points where $\omega$ is tangent to lines $B C, C A$, and $A B$, respectively. The circle $A E F$ intersects line $B C$ at $P$ and $Q$. Let $M$ be the midpoint of $A D$. Prove that the circle $M P Q$ is tangent to $\omega$.
(Denmark)
G5. Let $A B C C_{1} B_{1} A_{1}$ be a convex hexagon such that $A B=B C$, and suppose that the line segments $A A_{1}, B B_{1}$, and $C C_{1}$ have the same perpendicular bisector. Let the diagonals $A C_{1}$ and $A_{1} C$ meet at $D$, and denote by $\omega$ the circle $A B C$. Let $\omega$ intersect the circle $A_{1} B C_{1}$ again at $E \neq B$. Prove that the lines $B B_{1}$ and $D E$ intersect on $\omega$.
(Ukraine)
G6. Let $n \geqslant 3$ be an integer. Two regular $n$-gons $\mathcal{A}$ and $\mathcal{B}$ are given in the plane. Prove that the vertices of $\mathcal{A}$ that lie inside $\mathcal{B}$ or on its boundary are consecutive.
(That is, prove that there exists a line separating those vertices of $\mathcal{A}$ that lie inside $\mathcal{B}$ or on its boundary from the other vertices of $\mathcal{A}$.)
(Czech Republic)
G7. A convex quadrilateral $A B C D$ has an inscribed circle with center $I$. Let $I_{a}, I_{b}, I_{c}$, and $I_{d}$ be the incenters of the triangles $D A B, A B C, B C D$, and $C D A$, respectively. Suppose that the common external tangents of the circles $A I_{b} I_{d}$ and $C I_{b} I_{d}$ meet at $X$, and the common external tangents of the circles $B I_{a} I_{c}$ and $D I_{a} I_{c}$ meet at $Y$. Prove that $\angle X I Y=90^{\circ}$.
(Kazakhstan)
G8. There are 2017 mutually external circles drawn on a blackboard, such that no two are tangent and no three share a common tangent. A tangent segment is a line segment that is a common tangent to two circles, starting at one tangent point and ending at the other one. Luciano is drawing tangent segments on the blackboard, one at a time, so that no tangent segment intersects any other circles or previously drawn tangent segments. Luciano keeps drawing tangent segments until no more can be drawn. Find all possible numbers of tangent segments when he stops drawing.

## Number Theory

N1. The sequence $a_{0}, a_{1}, a_{2}, \ldots$ of positive integers satisfies

$$
a_{n+1}=\left\{\begin{array}{ll}
\sqrt{a_{n}}, & \text { if } \sqrt{a_{n}} \text { is an integer } \\
a_{n}+3, & \text { otherwise }
\end{array} \quad \text { for every } n \geqslant 0\right.
$$

Determine all values of $a_{0}>1$ for which there is at least one number $a$ such that $a_{n}=a$ for infinitely many values of $n$.
(South Africa)
N2. Let $p \geqslant 2$ be a prime number. Eduardo and Fernando play the following game making moves alternately: in each move, the current player chooses an index $i$ in the set $\{0,1, \ldots, p-1\}$ that was not chosen before by either of the two players and then chooses an element $a_{i}$ of the set $\{0,1,2,3,4,5,6,7,8,9\}$. Eduardo has the first move. The game ends after all the indices $i \in\{0,1, \ldots, p-1\}$ have been chosen. Then the following number is computed:

$$
M=a_{0}+10 \cdot a_{1}+\cdots+10^{p-1} \cdot a_{p-1}=\sum_{j=0}^{p-1} a_{j} \cdot 10^{j}
$$

The goal of Eduardo is to make the number $M$ divisible by $p$, and the goal of Fernando is to prevent this.

Prove that Eduardo has a winning strategy.
(Morocco)
N3. Determine all integers $n \geqslant 2$ with the following property: for any integers $a_{1}, a_{2}, \ldots, a_{n}$ whose sum is not divisible by $n$, there exists an index $1 \leqslant i \leqslant n$ such that none of the numbers

$$
a_{i}, a_{i}+a_{i+1}, \ldots, a_{i}+a_{i+1}+\cdots+a_{i+n-1}
$$

is divisible by $n$. (We let $a_{i}=a_{i-n}$ when $i>n$.)
(Thailand)
N4. Call a rational number short if it has finitely many digits in its decimal expansion. For a positive integer $m$, we say that a positive integer $t$ is $m$-tastic if there exists a number $c \in\{1,2,3, \ldots, 2017\}$ such that $\frac{10^{t}-1}{c \cdot m}$ is short, and such that $\frac{10^{k}-1}{c \cdot m}$ is not short for any $1 \leqslant k<t$. Let $S(m)$ be the set of $m$-tastic numbers. Consider $S(m)$ for $m=1,2, \ldots$. What is the maximum number of elements in $S(m)$ ?
(Turkey)
N5. Find all pairs $(p, q)$ of prime numbers with $p>q$ for which the number

$$
\frac{(p+q)^{p+q}(p-q)^{p-q}-1}{(p+q)^{p-q}(p-q)^{p+q}-1}
$$

is an integer.

N6. Find the smallest positive integer $n$, or show that no such $n$ exists, with the following property: there are infinitely many distinct $n$-tuples of positive rational numbers ( $a_{1}, a_{2}, \ldots, a_{n}$ ) such that both

$$
a_{1}+a_{2}+\cdots+a_{n} \quad \text { and } \quad \frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}
$$

are integers.
(Singapore)
N7. Say that an ordered pair $(x, y)$ of integers is an irreducible lattice point if $x$ and $y$ are relatively prime. For any finite set $S$ of irreducible lattice points, show that there is a homogenous polynomial in two variables, $f(x, y)$, with integer coefficients, of degree at least 1 , such that $f(x, y)=1$ for each $(x, y)$ in the set $S$.

Note: A homogenous polynomial of degree $n$ is any nonzero polynomial of the form

$$
\begin{equation*}
f(x, y)=a_{0} x^{n}+a_{1} x^{n-1} y+a_{2} x^{n-2} y^{2}+\cdots+a_{n-1} x y^{n-1}+a_{n} y^{n} . \tag{U.S.A.}
\end{equation*}
$$

N8. Let $p$ be an odd prime number and $\mathbb{Z}_{>0}$ be the set of positive integers. Suppose that a function $f: \mathbb{Z}_{>0} \times \mathbb{Z}_{>0} \rightarrow\{0,1\}$ satisfies the following properties:

- $f(1,1)=0$;
- $f(a, b)+f(b, a)=1$ for any pair of relatively prime positive integers $(a, b)$ not both equal to 1 ;
- $f(a+b, b)=f(a, b)$ for any pair of relatively prime positive integers $(a, b)$.

Prove that

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right) \geqslant \sqrt{2 p}-2
$$

## Solutions

## Algebra

A1. Let $a_{1}, a_{2}, \ldots, a_{n}, k$, and $M$ be positive integers such that

$$
\frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}=k \quad \text { and } \quad a_{1} a_{2} \ldots a_{n}=M
$$

If $M>1$, prove that the polynomial

$$
P(x)=M(x+1)^{k}-\left(x+a_{1}\right)\left(x+a_{2}\right) \cdots\left(x+a_{n}\right)
$$

has no positive roots.
(Trinidad and Tobago)
Solution 1. We first prove that, for $x>0$,

$$
\begin{equation*}
a_{i}(x+1)^{1 / a_{i}} \leqslant x+a_{i}, \tag{1}
\end{equation*}
$$

with equality if and only if $a_{i}=1$. It is clear that equality occurs if $a_{i}=1$.
If $a_{i}>1$, the AM-GM inequality applied to a single copy of $x+1$ and $a_{i}-1$ copies of 1 yields

$$
\frac{(x+1)+\overbrace{1+1+\cdots+1}^{a_{i}-1 \text { ones }}}{a_{i}} \geqslant \sqrt[a_{i}]{(x+1) \cdot 1^{a_{i}-1}} \Longrightarrow a_{i}(x+1)^{1 / a_{i}} \leqslant x+a_{i} .
$$

Since $x+1>1$, the inequality is strict for $a_{i}>1$.
Multiplying the inequalities (1) for $i=1,2, \ldots, n$ yields

$$
\prod_{i=1}^{n} a_{i}(x+1)^{1 / a_{i}} \leqslant \prod_{i=1}^{n}\left(x+a_{i}\right) \Longleftrightarrow M(x+1)^{\sum_{i=1}^{n} 1 / a_{i}}-\prod_{i=1}^{n}\left(x+a_{i}\right) \leqslant 0 \Longleftrightarrow P(x) \leqslant 0
$$

with equality iff $a_{i}=1$ for all $i \in\{1,2, \ldots, n\}$. But this implies $M=1$, which is not possible. Hence $P(x)<0$ for all $x \in \mathbb{R}^{+}$, and $P$ has no positive roots.

Comment 1. Inequality (1) can be obtained in several ways. For instance, we may also use the binomial theorem: since $a_{i} \geqslant 1$,

$$
\left(1+\frac{x}{a_{i}}\right)^{a_{i}}=\sum_{j=0}^{a_{i}}\binom{a_{i}}{j}\left(\frac{x}{a_{i}}\right)^{j} \geqslant\binom{ a_{i}}{0}+\binom{a_{i}}{1} \cdot \frac{x}{a_{i}}=1+x .
$$

Both proofs of (1) mimic proofs to Bernoulli's inequality for a positive integer exponent $a_{i}$; we can use this inequality directly:

$$
\left(1+\frac{x}{a_{i}}\right)^{a_{i}} \geqslant 1+a_{i} \cdot \frac{x}{a_{i}}=1+x,
$$

and so

$$
x+a_{i}=a_{i}\left(1+\frac{x}{a_{i}}\right) \geqslant a_{i}(1+x)^{1 / a_{i}},
$$

or its (reversed) formulation, with exponent $1 / a_{i} \leqslant 1$ :

$$
(1+x)^{1 / a_{i}} \leqslant 1+\frac{1}{a_{i}} \cdot x=\frac{x+a_{i}}{a_{i}} \Longrightarrow a_{i}(1+x)^{1 / a_{i}} \leqslant x+a_{i} .
$$

Solution 2. We will prove that, in fact, all coefficients of the polynomial $P(x)$ are non-positive, and at least one of them is negative, which implies that $P(x)<0$ for $x>0$.

Indeed, since $a_{j} \geqslant 1$ for all $j$ and $a_{j}>1$ for some $j$ (since $a_{1} a_{2} \ldots a_{n}=M>1$ ), we have $k=\frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}<n$, so the coefficient of $x^{n}$ in $P(x)$ is $-1<0$. Moreover, the coefficient of $x^{r}$ in $P(x)$ is negative for $k<r \leqslant n=\operatorname{deg}(P)$.

For $0 \leqslant r \leqslant k$, the coefficient of $x^{r}$ in $P(x)$ is

$$
M \cdot\binom{k}{r}-\sum_{1 \leqslant i_{1}<i_{2}<\cdots<i_{n-r} \leqslant n} a_{i_{1}} a_{i_{2}} \cdots a_{i_{n-r}}=a_{1} a_{2} \cdots a_{n} \cdot\binom{k}{r}-\sum_{1 \leqslant i_{1}<i_{2}<\cdots<i_{n-r} \leqslant n} a_{i_{1}} a_{i_{2}} \cdots a_{i_{n-r}},
$$

which is non-positive iff

$$
\begin{equation*}
\binom{k}{r} \leqslant \sum_{1 \leqslant j_{1}<j_{2}<\cdots<j_{r} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}} \tag{2}
\end{equation*}
$$

We will prove (2) by induction on $r$. For $r=0$ it is an equality because the constant term of $P(x)$ is $P(0)=0$, and if $r=1$, (2) becomes $k=\sum_{i=1}^{n} \frac{1}{a_{i}}$. For $r>1$, if (2) is true for a given $r<k$, we have

$$
\binom{k}{r+1}=\frac{k-r}{r+1} \cdot\binom{k}{r} \leqslant \frac{k-r}{r+1} . \sum_{1 \leqslant j_{1}<j_{2}<\cdots<j_{r} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}},
$$

and it suffices to prove that

$$
\frac{k-r}{r+1} \cdot \sum_{1 \leqslant j_{1}<j_{2}<\cdots<j_{r} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}} \leqslant \sum_{1 \leqslant j_{1}<\cdots<j_{r}<j_{r+1} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}} a_{j_{r+1}}},
$$

which is equivalent to

$$
\left(\frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}-r\right) \sum_{1 \leqslant j_{1}<j_{2}<\cdots<j_{r} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}} \leqslant(r+1) \sum_{1 \leqslant j_{1}<\cdots<j_{r}<j_{r+1} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}} a_{j_{r+1}}} .
$$

Since there are $r+1$ ways to choose a fraction $\frac{1}{a_{j_{i}}}$ from $\frac{1}{a_{j_{1} a_{j_{2}} \cdots a_{j_{r}} a_{j_{r}+1}}}$ to factor out, every term $\frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}} a_{j_{r+1}}}$ in the right hand side appears exactly $r+1$ times in the product

$$
\left(\frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}\right)_{1 \leqslant j_{1}<j_{2}<\cdots<j_{r} \leqslant n} \frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}} .
$$

Hence all terms in the right hand side cancel out.
The remaining terms in the left hand side can be grouped in sums of the type

$$
\begin{aligned}
\frac{1}{a_{j_{1}}^{2} a_{j_{2}} \cdots a_{j_{r}}}+\frac{1}{a_{j_{1}} a_{j_{2}}^{2} \cdots a_{j_{r}}}+\cdots+\frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}^{2}} & -\frac{r}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}} \\
& =\frac{1}{a_{j_{1}} a_{j_{2}} \cdots a_{j_{r}}}\left(\frac{1}{a_{j_{1}}}+\frac{1}{a_{j_{2}}}+\cdots+\frac{1}{a_{j_{r}}}-r\right),
\end{aligned}
$$

which are all non-positive because $a_{i} \geqslant 1 \Longrightarrow \frac{1}{a_{i}} \leqslant 1, i=1,2, \ldots, n$.
Comment 2. The result is valid for any real numbers $a_{i}, i=1,2, \ldots, n$ with $a_{i} \geqslant 1$ and product $M$ greater than 1. A variation of Solution 1, namely using weighted AM-GM (or the Bernoulli inequality for real exponents), actually proves that $P(x)<0$ for $x>-1$ and $x \neq 0$.

A2. Let $q$ be a real number. Gugu has a napkin with ten distinct real numbers written on it, and he writes the following three lines of real numbers on the blackboard:

- In the first line, Gugu writes down every number of the form $a-b$, where $a$ and $b$ are two (not necessarily distinct) numbers on his napkin.
- In the second line, Gugu writes down every number of the form $q a b$, where $a$ and $b$ are two (not necessarily distinct) numbers from the first line.
- In the third line, Gugu writes down every number of the form $a^{2}+b^{2}-c^{2}-d^{2}$, where $a, b, c, d$ are four (not necessarily distinct) numbers from the first line.

Determine all values of $q$ such that, regardless of the numbers on Gugu's napkin, every number in the second line is also a number in the third line.
(Austria)
Answer: - 2, 0, 2 .
Solution 1. Call a number $q$ good if every number in the second line appears in the third line unconditionally. We first show that the numbers 0 and $\pm 2$ are good. The third line necessarily contains 0 , so 0 is good. For any two numbers $a, b$ in the first line, write $a=x-y$ and $b=u-v$, where $x, y, u, v$ are (not necessarily distinct) numbers on the napkin. We may now write

$$
2 a b=2(x-y)(u-v)=(x-v)^{2}+(y-u)^{2}-(x-u)^{2}-(y-v)^{2},
$$

which shows that 2 is good. By negating both sides of the above equation, we also see that -2 is good.

We now show that $-2,0$, and 2 are the only good numbers. Assume for sake of contradiction that $q$ is a good number, where $q \notin\{-2,0,2\}$. We now consider some particular choices of numbers on Gugu's napkin to arrive at a contradiction.

Assume that the napkin contains the integers $1,2, \ldots, 10$. Then, the first line contains the integers $-9,-8, \ldots, 9$. The second line then contains $q$ and $81 q$, so the third line must also contain both of them. But the third line only contains integers, so $q$ must be an integer. Furthermore, the third line contains no number greater than $162=9^{2}+9^{2}-0^{2}-0^{2}$ or less than -162 , so we must have $-162 \leqslant 81 q \leqslant 162$. This shows that the only possibilities for $q$ are $\pm 1$.

Now assume that $q= \pm 1$. Let the napkin contain $0,1,4,8,12,16,20,24,28,32$. The first line contains $\pm 1$ and $\pm 4$, so the second line contains $\pm 4$. However, for every number $a$ in the first line, $a \not \equiv 2(\bmod 4)$, so we may conclude that $a^{2} \equiv 0,1(\bmod 8)$. Consequently, every number in the third line must be congruent to $-2,-1,0,1,2(\bmod 8)$; in particular, $\pm 4$ cannot be in the third line, which is a contradiction.

Solution 2. Let $q$ be a good number, as defined in the first solution, and define the polynomial $P\left(x_{1}, \ldots, x_{10}\right)$ as

$$
\prod_{i<j}\left(x_{i}-x_{j}\right) \prod_{a_{i} \in S}\left(q\left(x_{1}-x_{2}\right)\left(x_{3}-x_{4}\right)-\left(a_{1}-a_{2}\right)^{2}-\left(a_{3}-a_{4}\right)^{2}+\left(a_{5}-a_{6}\right)^{2}+\left(a_{7}-a_{8}\right)^{2}\right)
$$

where $S=\left\{x_{1}, \ldots, x_{10}\right\}$.
We claim that $P\left(x_{1}, \ldots, x_{10}\right)=0$ for every choice of real numbers $\left(x_{1}, \ldots, x_{10}\right)$. If any two of the $x_{i}$ are equal, then $P\left(x_{1}, \ldots, x_{10}\right)=0$ trivially. If no two are equal, assume that Gugu has those ten numbers $x_{1}, \ldots, x_{10}$ on his napkin. Then, the number $q\left(x_{1}-x_{2}\right)\left(x_{3}-x_{4}\right)$ is in the second line, so we must have some $a_{1}, \ldots, a_{8}$ so that

$$
q\left(x_{1}-x_{2}\right)\left(x_{3}-x_{4}\right)-\left(a_{1}-a_{2}\right)^{2}-\left(a_{3}-a_{4}\right)^{2}+\left(a_{5}-a_{6}\right)^{2}+\left(a_{7}-a_{8}\right)^{2}=0
$$

and hence $P\left(x_{1}, \ldots, x_{10}\right)=0$.
Since every polynomial that evaluates to zero everywhere is the zero polynomial, and the product of two nonzero polynomials is necessarily nonzero, we may define $F$ such that

$$
\begin{equation*}
F\left(x_{1}, \ldots, x_{10}\right) \equiv q\left(x_{1}-x_{2}\right)\left(x_{3}-x_{4}\right)-\left(a_{1}-a_{2}\right)^{2}-\left(a_{3}-a_{4}\right)^{2}+\left(a_{5}-a_{6}\right)^{2}+\left(a_{7}-a_{8}\right)^{2} \equiv 0 \tag{1}
\end{equation*}
$$

for some particular choice $a_{i} \in S$.
Each of the sets $\left\{a_{1}, a_{2}\right\},\left\{a_{3}, a_{4}\right\},\left\{a_{5}, a_{6}\right\}$, and $\left\{a_{7}, a_{8}\right\}$ is equal to at most one of the four sets $\left\{x_{1}, x_{3}\right\},\left\{x_{2}, x_{3}\right\},\left\{x_{1}, x_{4}\right\}$, and $\left\{x_{2}, x_{4}\right\}$. Thus, without loss of generality, we may assume that at most one of the sets $\left\{a_{1}, a_{2}\right\},\left\{a_{3}, a_{4}\right\},\left\{a_{5}, a_{6}\right\}$, and $\left\{a_{7}, a_{8}\right\}$ is equal to $\left\{x_{1}, x_{3}\right\}$. Let $u_{1}, u_{3}, u_{5}, u_{7}$ be the indicator functions for this equality of sets: that is, $u_{i}=1$ if and only if $\left\{a_{i}, a_{i+1}\right\}=\left\{x_{1}, x_{3}\right\}$. By assumption, at least three of the $u_{i}$ are equal to 0 .

We now compute the coefficient of $x_{1} x_{3}$ in $F$. It is equal to $q+2\left(u_{1}+u_{3}-u_{5}-u_{7}\right)=0$, and since at least three of the $u_{i}$ are zero, we must have that $q \in\{-2,0,2\}$, as desired.

A3. Let $S$ be a finite set, and let $\mathcal{A}$ be the set of all functions from $S$ to $S$. Let $f$ be an element of $\mathcal{A}$, and let $T=f(S)$ be the image of $S$ under $f$. Suppose that $f \circ g \circ f \neq g \circ f \circ g$ for every $g$ in $\mathcal{A}$ with $g \neq f$. Show that $f(T)=T$.
(India)
Solution. For $n \geqslant 1$, denote the $n$-th composition of $f$ with itself by

$$
f^{n} \stackrel{\text { def }}{=} \underbrace{f \circ f \circ \cdots \circ f}_{n \text { times }} .
$$

By hypothesis, if $g \in \mathcal{A}$ satisfies $f \circ g \circ f=g \circ f \circ g$, then $g=f$. A natural idea is to try to plug in $g=f^{n}$ for some $n$ in the expression $f \circ g \circ f=g \circ f \circ g$ in order to get $f^{n}=f$, which solves the problem:
Claim. If there exists $n \geqslant 3$ such that $f^{n+2}=f^{2 n+1}$, then the restriction $f: T \rightarrow T$ of $f$ to $T$ is a bijection.
Proof. Indeed, by hypothesis, $f^{n+2}=f^{2 n+1} \Longleftrightarrow f \circ f^{n} \circ f=f^{n} \circ f \circ f^{n} \Longrightarrow f^{n}=f$. Since $n-2 \geqslant 1$, the image of $f^{n-2}$ is contained in $T=f(S)$, hence $f^{n-2}$ restricts to a function $f^{n-2}: T \rightarrow T$. This is the inverse of $f: T \rightarrow T$. In fact, given $t \in T$, say $t=f(s)$ with $s \in S$, we have

$$
t=f(s)=f^{n}(s)=f^{n-2}(f(t))=f\left(f^{n-2}(t)\right), \quad \text { i.e., } \quad f^{n-2} \circ f=f \circ f^{n-2}=\text { id on } T
$$

(here id stands for the identity function). Hence, the restriction $f: T \rightarrow T$ of $f$ to $T$ is bijective with inverse given by $f^{n-2}: T \rightarrow T$.

It remains to show that $n$ as in the claim exists. For that, define

$$
S_{m} \stackrel{\text { def }}{=} f^{m}(S) \quad\left(S_{m} \text { is image of } f^{m}\right)
$$

Clearly the image of $f^{m+1}$ is contained in the image of $f^{m}$, i.e., there is a descending chain of subsets of $S$

$$
S \supseteq S_{1} \supseteq S_{2} \supseteq S_{3} \supseteq S_{4} \supseteq \cdots,
$$

which must eventually stabilise since $S$ is finite, i.e., there is a $k \geqslant 1$ such that

$$
S_{k}=S_{k+1}=S_{k+2}=S_{k+3}=\cdots \stackrel{\text { def }}{=} S_{\infty} .
$$

Hence $f$ restricts to a surjective function $f: S_{\infty} \rightarrow S_{\infty}$, which is also bijective since $S_{\infty} \subseteq S$ is finite. To sum up, $f: S_{\infty} \rightarrow S_{\infty}$ is a permutation of the elements of the finite set $S_{\infty}$, hence there exists an integer $r \geqslant 1$ such that $f^{r}=\mathrm{id}$ on $S_{\infty}$ (for example, we may choose $r=\left|S_{\infty}\right|!$ ). In other words,

$$
\begin{equation*}
f^{m+r}=f^{m} \text { on } S \text { for all } m \geqslant k . \tag{*}
\end{equation*}
$$

Clearly, (*) also implies that $f^{m+t r}=f^{m}$ for all integers $t \geqslant 1$ and $m \geqslant k$. So, to find $n$ as in the claim and finish the problem, it is enough to choose $m$ and $t$ in order to ensure that there exists $n \geqslant 3$ satisfying

$$
\left\{\begin{array} { l } 
{ 2 n + 1 = m + t r } \\
{ n + 2 = m }
\end{array} \Longleftrightarrow \left\{\begin{array}{l}
m=3+t r \\
n=m-2
\end{array}\right.\right.
$$

This can be clearly done by choosing $m$ large enough with $m \equiv 3(\bmod r)$. For instance, we may take $n=2 k r+1$, so that

$$
f^{n+2}=f^{2 k r+3}=f^{4 k r+3}=f^{2 n+1}
$$

where the middle equality follows by (*) since $2 k r+3 \geqslant k$.

A4. A sequence of real numbers $a_{1}, a_{2}, \ldots$ satisfies the relation

$$
a_{n}=-\max _{i+j=n}\left(a_{i}+a_{j}\right) \quad \text { for all } n>2017
$$

Prove that this sequence is bounded, i.e., there is a constant $M$ such that $\left|a_{n}\right| \leqslant M$ for all positive integers $n$.
(Russia)
Solution 1. Set $D=2017$. Denote

$$
M_{n}=\max _{k<n} a_{k} \quad \text { and } \quad m_{n}=-\min _{k<n} a_{k}=\max _{k<n}\left(-a_{k}\right) .
$$

Clearly, the sequences $\left(m_{n}\right)$ and $\left(M_{n}\right)$ are nondecreasing. We need to prove that both are bounded.

Consider an arbitrary $n>D$; our first aim is to bound $a_{n}$ in terms of $m_{n}$ and $M_{n}$.
(i) There exist indices $p$ and $q$ such that $a_{n}=-\left(a_{p}+a_{q}\right)$ and $p+q=n$. Since $a_{p}, a_{q} \leqslant M_{n}$, we have $a_{n} \geqslant-2 M_{n}$.
(ii) On the other hand, choose an index $k<n$ such that $a_{k}=M_{n}$. Then, we have

$$
a_{n}=-\max _{\ell<n}\left(a_{n-\ell}+a_{\ell}\right) \leqslant-\left(a_{n-k}+a_{k}\right)=-a_{n-k}-M_{n} \leqslant m_{n}-M_{n} .
$$

Summarizing (i) and (ii), we get

$$
-2 M_{n} \leqslant a_{n} \leqslant m_{n}-M_{n},
$$

whence

$$
\begin{equation*}
m_{n} \leqslant m_{n+1} \leqslant \max \left\{m_{n}, 2 M_{n}\right\} \quad \text { and } \quad M_{n} \leqslant M_{n+1} \leqslant \max \left\{M_{n}, m_{n}-M_{n}\right\} \tag{1}
\end{equation*}
$$

Now, say that an index $n>D$ is lucky if $m_{n} \leqslant 2 M_{n}$. Two cases are possible.
Case 1. Assume that there exists a lucky index $n$. In this case, (1) yields $m_{n+1} \leqslant 2 M_{n}$ and $M_{n} \leqslant M_{n+1} \leqslant M_{n}$. Therefore, $M_{n+1}=M_{n}$ and $m_{n+1} \leqslant 2 M_{n}=2 M_{n+1}$. So, the index $n+1$ is also lucky, and $M_{n+1}=M_{n}$. Applying the same arguments repeatedly, we obtain that all indices $k>n$ are lucky (i.e., $m_{k} \leqslant 2 M_{k}$ for all these indices), and $M_{k}=M_{n}$ for all such indices. Thus, all of the $m_{k}$ and $M_{k}$ are bounded by $2 M_{n}$.
Case 2. Assume now that there is no lucky index, i.e., $2 M_{n}<m_{n}$ for all $n>D$. Then (1) shows that for all $n>D$ we have $m_{n} \leqslant m_{n+1} \leqslant m_{n}$, so $m_{n}=m_{D+1}$ for all $n>D$. Since $M_{n}<m_{n} / 2$ for all such indices, all of the $m_{n}$ and $M_{n}$ are bounded by $m_{D+1}$.

Thus, in both cases the sequences $\left(m_{n}\right)$ and $\left(M_{n}\right)$ are bounded, as desired.
Solution 2. As in the previous solution, let $D=2017$. If the sequence is bounded above, say, by $Q$, then we have that $a_{n} \geqslant \min \left\{a_{1}, \ldots, a_{D},-2 Q\right\}$ for all $n$, so the sequence is bounded. Assume for sake of contradiction that the sequence is not bounded above. Let $\ell=\min \left\{a_{1}, \ldots, a_{D}\right\}$, and $L=\max \left\{a_{1}, \ldots, a_{D}\right\}$. Call an index $n$ good if the following criteria hold:

$$
\begin{equation*}
a_{n}>a_{i} \text { for each } i<n, \quad a_{n}>-2 \ell, \quad \text { and } \quad n>D \tag{2}
\end{equation*}
$$

We first show that there must be some good index $n$. By assumption, we may take an index $N$ such that $a_{N}>\max \{L,-2 \ell\}$. Choose $n$ minimally such that $a_{n}=\max \left\{a_{1}, a_{2}, \ldots, a_{N}\right\}$. Now, the first condition in (2) is satisfied because of the minimality of $n$, and the second and third conditions are satisfied because $a_{n} \geqslant a_{N}>L,-2 \ell$, and $L \geqslant a_{i}$ for every $i$ such that $1 \leqslant i \leqslant D$.

Let $n$ be a good index. We derive a contradiction. We have that

$$
\begin{equation*}
a_{n}+a_{u}+a_{v} \leqslant 0, \tag{3}
\end{equation*}
$$

whenever $u+v=n$.
We define the index $u$ to maximize $a_{u}$ over $1 \leqslant u \leqslant n-1$, and let $v=n-u$. Then, we note that $a_{u} \geqslant a_{v}$ by the maximality of $a_{u}$.

Assume first that $v \leqslant D$. Then, we have that

$$
a_{N}+2 \ell \leqslant 0,
$$

because $a_{u} \geqslant a_{v} \geqslant \ell$. But this contradicts our assumption that $a_{n}>-2 \ell$ in the second criteria of (2).

Now assume that $v>D$. Then, there exist some indices $w_{1}, w_{2}$ summing up to $v$ such that

$$
a_{v}+a_{w_{1}}+a_{w_{2}}=0 .
$$

But combining this with (3), we have

$$
a_{n}+a_{u} \leqslant a_{w_{1}}+a_{w_{2}} .
$$

Because $a_{n}>a_{u}$, we have that $\max \left\{a_{w_{1}}, a_{w_{2}}\right\}>a_{u}$. But since each of the $w_{i}$ is less than $v$, this contradicts the maximality of $a_{u}$.

Comment 1. We present two harder versions of this problem below.
Version 1. Let $a_{1}, a_{2}, \ldots$ be a sequence of numbers that satisfies the relation

$$
a_{n}=-\max _{i+j+k=n}\left(a_{i}+a_{j}+a_{k}\right) \quad \text { for all } n>2017 .
$$

Then, this sequence is bounded.
Proof. Set $D=2017$. Denote

$$
M_{n}=\max _{k<n} a_{k} \quad \text { and } \quad m_{n}=-\min _{k<n} a_{k}=\max _{k<n}\left(-a_{k}\right) .
$$

Clearly, the sequences $\left(m_{n}\right)$ and $\left(M_{n}\right)$ are nondecreasing. We need to prove that both are bounded.
Consider an arbitrary $n>2 D$; our first aim is to bound $a_{n}$ in terms of $m_{i}$ and $M_{i}$. Set $k=\lfloor n / 2\rfloor$.
(i) Choose indices $p, q$, and $r$ such that $a_{n}=-\left(a_{p}+a_{q}+a_{r}\right)$ and $p+q+r=n$. Without loss of generality, $p \geqslant q \geqslant r$.

Assume that $p \geqslant k+1(>D)$; then $p>q+r$. Hence

$$
-a_{p}=\max _{i_{1}+i_{2}+i_{3}=p}\left(a_{i_{1}}+a_{i_{2}}+a_{i_{3}}\right) \geqslant a_{q}+a_{r}+a_{p-q-r},
$$

and therefore $a_{n}=-\left(a_{p}+a_{q}+a_{r}\right) \geqslant\left(a_{q}+a_{r}+a_{p-q-r}\right)-a_{q}-a_{r}=a_{p-q-r} \geqslant-m_{n}$.
Otherwise, we have $k \geqslant p \geqslant q \geqslant r$. Since $n<3 k$, we have $r<k$. Then $a_{p}, a_{q} \leqslant M_{k+1}$ and $a_{r} \leqslant M_{k}$, whence $a_{n} \geqslant-2 M_{k+1}-M_{k}$.

Thus, in any case $a_{n} \geqslant-\max \left\{m_{n}, 2 M_{k+1}+M_{k}\right\}$.
(ii) On the other hand, choose $p \leqslant k$ and $q \leqslant k-1$ such that $a_{p}=M_{k+1}$ and $a_{q}=M_{k}$. Then $p+q<n$, so $a_{n} \leqslant-\left(a_{p}+a_{q}+a_{n-p-q}\right)=-a_{n-p-q}-M_{k+1}-M_{k} \leqslant m_{n}-M_{k+1}-M_{k}$.

To summarize,

$$
-\max \left\{m_{n}, 2 M_{k+1}+M_{k}\right\} \leqslant a_{n} \leqslant m_{n}-M_{k+1}-M_{k},
$$

whence

$$
\begin{equation*}
m_{n} \leqslant m_{n+1} \leqslant \max \left\{m_{n}, 2 M_{k+1}+M_{k}\right\} \quad \text { and } \quad M_{n} \leqslant M_{n+1} \leqslant \max \left\{M_{n}, m_{n}-M_{k+1}-M_{k}\right\} . \tag{4}
\end{equation*}
$$

Now, say that an index $n>2 D$ is lucky if $m_{n} \leqslant 2 M_{\lfloor n / 2\rfloor+1}+M_{[n / 2]}$. Two cases are possible.
Case 1. Assume that there exists a lucky index $n$; set $k=\lfloor n / 2\rfloor$. In this case, (4) yields $m_{n+1} \leqslant$ $2 M_{k+1}+M_{k}$ and $M_{n} \leqslant M_{n+1} \leqslant M_{n}$ (the last relation holds, since $m_{n}-M_{k+1}-M_{k} \leqslant\left(2 M_{k+1}+\right.$ $\left.M_{k}\right)-M_{k+1}-M_{k}=M_{k+1} \leqslant M_{n}$ ). Therefore, $M_{n+1}=M_{n}$ and $m_{n+1} \leqslant 2 M_{k+1}+M_{k}$; the last relation shows that the index $n+1$ is also lucky.

Thus, all indices $N>n$ are lucky, and $M_{N}=M_{n} \geqslant m_{N} / 3$, whence all the $m_{N}$ and $M_{N}$ are bounded by $3 M_{n}$.
Case 2. Conversely, assume that there is no lucky index, i.e., $2 M_{[n / 2]+1}+M_{\lfloor n / 2]}<m_{n}$ for all $n>2 D$. Then (4) shows that for all $n>2 D$ we have $m_{n} \leqslant m_{n+1} \leqslant m_{n}$, i.e., $m_{N}=m_{2 D+1}$ for all $N>2 D$. Since $M_{N}<m_{2 N+1} / 3$ for all such indices, all the $m_{N}$ and $M_{N}$ are bounded by $m_{2 D+1}$.

Thus, in both cases the sequences $\left(m_{n}\right)$ and $\left(M_{n}\right)$ are bounded, as desired.
Version 2. Let $a_{1}, a_{2}, \ldots$ be a sequence of numbers that satisfies the relation

$$
a_{n}=-\max _{i_{1}+\cdots+i_{k}=n}\left(a_{i_{1}}+\cdots+a_{i_{k}}\right) \quad \text { for all } n>2017 .
$$

Then, this sequence is bounded.
Proof. As in the solutions above, let $D=2017$. If the sequence is bounded above, say, by $Q$, then we have that $a_{n} \geqslant \min \left\{a_{1}, \ldots, a_{D},-k Q\right\}$ for all $n$, so the sequence is bounded. Assume for sake of contradiction that the sequence is not bounded above. Let $\ell=\min \left\{a_{1}, \ldots, a_{D}\right\}$, and $L=\max \left\{a_{1}, \ldots, a_{D}\right\}$. Call an index $n$ good if the following criteria hold:

$$
\begin{equation*}
a_{n}>a_{i} \text { for each } i<n, \quad a_{n}>-k \ell, \quad \text { and } \quad n>D \tag{5}
\end{equation*}
$$

We first show that there must be some good index $n$. By assumption, we may take an index $N$ such that $a_{N}>\max \{L,-k \ell\}$. Choose $n$ minimally such that $a_{n}=\max \left\{a_{1}, a_{2}, \ldots, a_{N}\right\}$. Now, the first condition is satisfied because of the minimality of $n$, and the second and third conditions are satisfied because $a_{n} \geqslant a_{N}>L,-k \ell$, and $L \geqslant a_{i}$ for every $i$ such that $1 \leqslant i \leqslant D$.

Let $n$ be a good index. We derive a contradiction. We have that

$$
\begin{equation*}
a_{n}+a_{v_{1}}+\cdots+a_{v_{k}} \leqslant 0, \tag{6}
\end{equation*}
$$

whenever $v_{1}+\cdots+v_{k}=n$.
We define the sequence of indices $v_{1}, \ldots, v_{k-1}$ to greedily maximize $a_{v_{1}}$, then $a_{v_{2}}$, and so forth, selecting only from indices such that the equation $v_{1}+\cdots+v_{k}=n$ can be satisfied by positive integers $v_{1}, \ldots, v_{k}$. More formally, we define them inductively so that the following criteria are satisfied by the $v_{i}$ :

1. $1 \leqslant v_{i} \leqslant n-(k-i)-\left(v_{1}+\cdots+v_{i-1}\right)$.
2. $a_{v_{i}}$ is maximal among all choices of $v_{i}$ from the first criteria.

First of all, we note that for each $i$, the first criteria is always satisfiable by some $v_{i}$, because we are guaranteed that

$$
v_{i-1} \leqslant n-(k-(i-1))-\left(v_{1}+\cdots+v_{i-2}\right),
$$

which implies

$$
1 \leqslant n-(k-i)-\left(v_{1}+\cdots+v_{i-1}\right) .
$$

Secondly, the sum $v_{1}+\cdots+v_{k-1}$ is at most $n-1$. Define $v_{k}=n-\left(v_{1}+\cdots+v_{k-1}\right)$. Then, (6) is satisfied by the $v_{i}$. We also note that $a_{v_{i}} \geqslant a_{v_{j}}$ for all $i<j$; otherwise, in the definition of $v_{i}$, we could have selected $v_{j}$ instead.

Assume first that $v_{k} \leqslant D$. Then, from (6), we have that

$$
a_{n}+k \ell \leqslant 0,
$$

by using that $a_{v_{1}} \geqslant \cdots \geqslant a_{v_{k}} \geqslant \ell$. But this contradicts our assumption that $a_{n}>-k \ell$ in the second criteria of (5).

Now assume that $v_{k}>D$, and then we must have some indices $w_{1}, \ldots, w_{k}$ summing up to $v_{k}$ such that

$$
a_{v_{k}}+a_{w_{1}}+\cdots+a_{w_{k}}=0 .
$$

But combining this with (6), we have

$$
a_{n}+a_{v_{1}}+\cdots+a_{v_{k-1}} \leqslant a_{w_{1}}+\cdots+a_{w_{k}} .
$$

Because $a_{n}>a_{v_{1}} \geqslant \cdots \geqslant a_{v_{k-1}}$, we have that $\max \left\{a_{w_{1}}, \ldots, a_{w_{k}}\right\}>a_{v_{k-1}}$. But since each of the $w_{i}$ is less than $v_{k}$, in the definition of the $v_{k-1}$ we could have chosen one of the $w_{i}$ instead, which is a contradiction.

Comment 2. It seems that each sequence satisfying the condition in Version 2 is eventually periodic, at least when its terms are integers.

However, up to this moment, the Problem Selection Committee is not aware of a proof for this fact (even in the case $k=2$ ).

A5. An integer $n \geqslant 3$ is given. We call an $n$-tuple of real numbers $\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ Shiny if for each permutation $y_{1}, y_{2}, \ldots, y_{n}$ of these numbers we have

$$
\sum_{i=1}^{n-1} y_{i} y_{i+1}=y_{1} y_{2}+y_{2} y_{3}+y_{3} y_{4}+\cdots+y_{n-1} y_{n} \geqslant-1 .
$$

Find the largest constant $K=K(n)$ such that

$$
\sum_{1 \leqslant i<j \leqslant n} x_{i} x_{j} \geqslant K
$$

holds for every Shiny $n$-tuple $\left(x_{1}, x_{2}, \ldots, x_{n}\right)$.
(Serbia)
Answer: $K=-(n-1) / 2$.
Solution 1. First of all, we show that we may not take a larger constant $K$. Let $t$ be a positive number, and take $x_{2}=x_{3}=\cdots=t$ and $x_{1}=-1 /(2 t)$. Then, every product $x_{i} x_{j}(i \neq j)$ is equal to either $t^{2}$ or $-1 / 2$. Hence, for every permutation $y_{i}$ of the $x_{i}$, we have

$$
y_{1} y_{2}+\cdots+y_{n-1} y_{n} \geqslant(n-3) t^{2}-1 \geqslant-1 .
$$

This justifies that the $n$-tuple $\left(x_{1}, \ldots, x_{n}\right)$ is Shiny. Now, we have

$$
\sum_{i<j} x_{i} x_{j}=-\frac{n-1}{2}+\frac{(n-1)(n-2)}{2} t^{2} .
$$

Thus, as $t$ approaches 0 from above, $\sum_{i<j} x_{i} x_{j}$ gets arbitrarily close to $-(n-1) / 2$. This shows that we may not take $K$ any larger than $-(n-1) / 2$. It remains to show that $\sum_{i<j} x_{i} x_{j} \geqslant$ $-(n-1) / 2$ for any Shiny choice of the $x_{i}$.

From now onward, assume that $\left(x_{1}, \ldots, x_{n}\right)$ is a Shiny $n$-tuple. Let the $z_{i}(1 \leqslant i \leqslant n)$ be some permutation of the $x_{i}$ to be chosen later. The indices for $z_{i}$ will always be taken modulo $n$. We will first split up the sum $\sum_{i<j} x_{i} x_{j}=\sum_{i<j} z_{i} z_{j}$ into $\lfloor(n-1) / 2\rfloor$ expressions, each of the form $y_{1} y_{2}+\cdots+y_{n-1} y_{n}$ for some permutation $y_{i}$ of the $z_{i}$, and some leftover terms. More specifically, write

$$
\begin{equation*}
\sum_{i<j} z_{i} z_{j}=\sum_{q=0}^{n-1} \sum_{\substack{i+j \equiv q \\ i \neq j}} z_{i} z_{j}=\sum_{p=1}^{\lfloor\bmod n)}, ~ \sum_{\substack{i+j \equiv 2 p-1,2 p \\ i \neq j \\(\bmod (\bmod n)}} z_{i} z_{j}+L, \tag{1}
\end{equation*}
$$

where $L=z_{1} z_{-1}+z_{2} z_{-2}+\cdots+z_{(n-1) / 2} z_{-(n-1) / 2}$ if $n$ is odd, and $L=z_{1} z_{-1}+z_{1} z_{-2}+z_{2} z_{-2}+$ $\cdots+z_{(n-2) / 2} z_{-n / 2}$ if $n$ is even. We note that for each $p=1,2, \ldots,\lfloor(n-1) / 2\rfloor$, there is some permutation $y_{i}$ of the $z_{i}$ such that

$$
\sum_{\substack{i+j \equiv 2 p-1,2 p \\ i \neq j \\(\bmod n)}} z_{i} z_{j}=\sum_{k=1}^{n-1} y_{k} y_{k+1},
$$

because we may choose $y_{2 i-1}=z_{i+p-1}$ for $1 \leqslant i \leqslant(n+1) / 2$ and $y_{2 i}=z_{p-i}$ for $1 \leqslant i \leqslant n / 2$.
We show (1) graphically for $n=6,7$ in the diagrams below. The edges of the graphs each represent a product $z_{i} z_{j}$, and the dashed and dotted series of lines represents the sum of the edges, which is of the form $y_{1} y_{2}+\cdots+y_{n-1} y_{n}$ for some permutation $y_{i}$ of the $z_{i}$ precisely when the series of lines is a Hamiltonian path. The filled edges represent the summands of $L$.


Now, because the $z_{i}$ are Shiny, we have that (1) yields the following bound:

$$
\sum_{i<j} z_{i} z_{j} \geqslant-\left\lfloor\frac{n-1}{2}\right\rfloor+L .
$$

It remains to show that, for each $n$, there exists some permutation $z_{i}$ of the $x_{i}$ such that $L \geqslant 0$ when $n$ is odd, and $L \geqslant-1 / 2$ when $n$ is even. We now split into cases based on the parity of $n$ and provide constructions of the permutations $z_{i}$.

Since we have not made any assumptions yet about the $x_{i}$, we may now assume without loss of generality that

$$
\begin{equation*}
x_{1} \leqslant x_{2} \leqslant \cdots \leqslant x_{k} \leqslant 0 \leqslant x_{k+1} \leqslant \cdots \leqslant x_{n} \tag{2}
\end{equation*}
$$

Case 1: $n$ is odd.
Without loss of generality, assume that $k$ (from (2)) is even, because we may negate all the $x_{i}$ if $k$ is odd. We then have $x_{1} x_{2}, x_{3} x_{4}, \ldots, x_{n-2} x_{n-1} \geqslant 0$ because the factors are of the same sign. Let $L=x_{1} x_{2}+x_{3} x_{4}+\cdots+x_{n-2} x_{n-1} \geqslant 0$. We choose our $z_{i}$ so that this definition of $L$ agrees with the sum of the leftover terms in (1). Relabel the $x_{i}$ as $z_{i}$ such that

$$
\left\{z_{1}, z_{n-1}\right\},\left\{z_{2}, z_{n-2}\right\}, \ldots,\left\{z_{(n-1) / 2}, z_{(n+1) / 2}\right\}
$$

are some permutation of

$$
\left\{x_{1}, x_{2}\right\},\left\{x_{3}, x_{4}\right\}, \ldots,\left\{x_{n-2}, x_{n-1}\right\}
$$

and $z_{n}=x_{n}$. Then, we have $L=z_{1} z_{n-1}+\cdots+z_{(n-1) / 2} z_{(n+1) / 2}$, as desired.
Case 2: $n$ is even.
Let $L=x_{1} x_{2}+x_{2} x_{3}+\cdots+x_{n-1} x_{n}$. Assume without loss of generality $k \neq 1$. Now, we have

$$
\begin{gathered}
2 L=\left(x_{1} x_{2}+\cdots+x_{n-1} x_{n}\right)+\left(x_{1} x_{2}+\cdots+x_{n-1} x_{n}\right) \geqslant\left(x_{2} x_{3}+\cdots+x_{n-1} x_{n}\right)+x_{k} x_{k+1} \\
\geqslant x_{2} x_{3}+\cdots+x_{n-1} x_{n}+x_{n} x_{1} \geqslant-1,
\end{gathered}
$$

where the first inequality holds because the only negative term in $L$ is $x_{k} x_{k+1}$, the second inequality holds because $x_{1} \leqslant x_{k} \leqslant 0 \leqslant x_{k+1} \leqslant x_{n}$, and the third inequality holds because the $x_{i}$ are assumed to be Shiny. We thus have that $L \geqslant-1 / 2$. We now choose a suitable $z_{i}$ such that the definition of $L$ matches the leftover terms in (1).

Relabel the $x_{i}$ with $z_{i}$ in the following manner: $x_{2 i-1}=z_{-i}, x_{2 i}=z_{i}$ (again taking indices modulo $n$ ). We have that

$$
L=\sum_{\substack{i+j \equiv 0,-1(\bmod n) \\ i \neq j}} z_{i} z_{j},
$$

as desired.
Solution 2. We present another proof that $\sum_{i<j} x_{i} x_{j} \geqslant-(n-1) / 2$ for any Shiny $n$-tuple $\left(x_{1}, \ldots, x_{n}\right)$. Assume an ordering of the $x_{i}$ as in (2), and let $\ell=n-k$. Assume without loss of generality that $k \geqslant \ell$. Also assume $k \neq n$, (as otherwise, all of the $x_{i}$ are nonpositive, and so the inequality is trivial). Define the sets of indices $S=\{1,2, \ldots, k\}$ and $T=\{k+1, \ldots, n\}$. Define the following sums:

$$
K=\sum_{\substack{i<j \\ i, j \in S}} x_{i} x_{j}, \quad M=\sum_{\substack{i \in S \\ j \in T}} x_{i} x_{j}, \quad \text { and } \quad L=\sum_{\substack{i<j \\ i, j \in T}} x_{i} x_{j}
$$

By definition, $K, L \geqslant 0$ and $M \leqslant 0$. We aim to show that $K+L+M \geqslant-(n-1) / 2$.
We split into cases based on whether $k=\ell$ or $k>\ell$.
Case 1: $k>\ell$.
Consider all permutations $\phi:\{1,2, \ldots, n\} \rightarrow\{1,2, \ldots, n\}$ such that $\phi^{-1}(T)=\{2,4, \ldots, 2 \ell\}$. Note that there are $k!!$ ! such permutations $\phi$. Define

$$
f(\phi)=\sum_{i=1}^{n-1} x_{\phi(i)} x_{\phi(i+1)} .
$$

We know that $f(\phi) \geqslant-1$ for every permutation $\phi$ with the above property. Averaging $f(\phi)$ over all $\phi$ gives

$$
-1 \leqslant \frac{1}{k!\ell!} \sum_{\phi} f(\phi)=\frac{2 \ell}{k \ell} M+\frac{2(k-\ell-1)}{k(k-1)} K
$$

where the equality holds because there are $k \ell$ products in $M$, of which $2 \ell$ are selected for each $\phi$, and there are $k(k-1) / 2$ products in $K$, of which $k-\ell-1$ are selected for each $\phi$. We now have

$$
K+L+M \geqslant K+L+\left(-\frac{k}{2}-\frac{k-\ell-1}{k-1} K\right)=-\frac{k}{2}+\frac{\ell}{k-1} K+L .
$$

Since $k \leqslant n-1$ and $K, L \geqslant 0$, we get the desired inequality.
Case 2: $k=\ell=n / 2$.
We do a similar approach, considering all $\phi:\{1,2, \ldots, n\} \rightarrow\{1,2, \ldots, n\}$ such that $\phi^{-1}(T)=$ $\{2,4, \ldots, 2 \ell\}$, and defining $f$ the same way. Analogously to Case 1 , we have

$$
-1 \leqslant \frac{1}{k!\ell!} \sum_{\phi} f(\phi)=\frac{2 \ell-1}{k \ell} M
$$

because there are $k \ell$ products in $M$, of which $2 \ell-1$ are selected for each $\phi$. Now, we have that

$$
K+L+M \geqslant M \geqslant-\frac{n^{2}}{4(n-1)} \geqslant-\frac{n-1}{2}
$$

where the last inequality holds because $n \geqslant 4$.

A6. Find all functions $f: \mathbb{R} \rightarrow \mathbb{R}$ such that

$$
\begin{equation*}
f(f(x) f(y))+f(x+y)=f(x y) \tag{*}
\end{equation*}
$$

for all $x, y \in \mathbb{R}$.
(Albania)
Answer: There are 3 solutions:

$$
x \mapsto 0 \quad \text { or } \quad x \mapsto x-1 \quad \text { or } \quad x \mapsto 1-x \quad(x \in \mathbb{R}) .
$$

Solution. An easy check shows that all the 3 above mentioned functions indeed satisfy the original equation (*).

In order to show that these are the only solutions, first observe that if $f(x)$ is a solution then $-f(x)$ is also a solution. Hence, without loss of generality we may (and will) assume that $f(0) \leqslant 0$ from now on. We have to show that either $f$ is identically zero or $f(x)=x-1$ $(\forall x \in \mathbb{R})$.

Observe that, for a fixed $x \neq 1$, we may choose $y \in \mathbb{R}$ so that $x+y=x y \Longleftrightarrow y=\frac{x}{x-1}$, and therefore from the original equation $(*)$ we have

$$
\begin{equation*}
f\left(f(x) \cdot f\left(\frac{x}{x-1}\right)\right)=0 \quad(x \neq 1) \tag{1}
\end{equation*}
$$

In particular, plugging in $x=0$ in (1), we conclude that $f$ has at least one zero, namely $(f(0))^{2}$ :

$$
\begin{equation*}
f\left((f(0))^{2}\right)=0 \tag{2}
\end{equation*}
$$

We analyze two cases (recall that $f(0) \leqslant 0$ ):
Case 1: $f(0)=0$.
Setting $y=0$ in the original equation we get the identically zero solution:

$$
f(f(x) f(0))+f(x)=f(0) \Longrightarrow f(x)=0 \text { for all } x \in \mathbb{R}
$$

From now on, we work on the main
Case 2: $f(0)<0$.
We begin with the following

## Claim 1.

$$
\begin{equation*}
f(1)=0, \quad f(a)=0 \Longrightarrow a=1, \quad \text { and } \quad f(0)=-1 . \tag{3}
\end{equation*}
$$

Proof. We need to show that 1 is the unique zero of $f$. First, observe that $f$ has at least one zero $a$ by (2); if $a \neq 1$ then setting $x=a$ in (1) we get $f(0)=0$, a contradiction. Hence from (2) we get $(f(0))^{2}=1$. Since we are assuming $f(0)<0$, we conclude that $f(0)=-1$.

Setting $y=1$ in the original equation (*) we get

$$
f(f(x) f(1))+f(x+1)=f(x) \Longleftrightarrow f(0)+f(x+1)=f(x) \Longleftrightarrow f(x+1)=f(x)+1 \quad(x \in \mathbb{R}) .
$$

An easy induction shows that

$$
\begin{equation*}
f(x+n)=f(x)+n \quad(x \in \mathbb{R}, n \in \mathbb{Z}) \tag{4}
\end{equation*}
$$

Now we make the following
Claim 2. $f$ is injective.
Proof. Suppose that $f(a)=f(b)$ with $a \neq b$. Then by (4), for all $N \in \mathbb{Z}$,

$$
f(a+N+1)=f(b+N)+1
$$

Choose any integer $N<-b$; then there exist $x_{0}, y_{0} \in \mathbb{R}$ with $x_{0}+y_{0}=a+N+1, x_{0} y_{0}=b+N$. Since $a \neq b$, we have $x_{0} \neq 1$ and $y_{0} \neq 1$. Plugging in $x_{0}$ and $y_{0}$ in the original equation (*) we get

$$
\begin{align*}
f\left(f\left(x_{0}\right) f\left(y_{0}\right)\right)+f(a+N+1)=f(b+N) & \Longleftrightarrow f\left(f\left(x_{0}\right) f\left(y_{0}\right)\right)+1=0 \\
& \Longleftrightarrow f\left(f\left(x_{0}\right) f\left(y_{0}\right)+1\right)=0  \tag{4}\\
& \Longleftrightarrow f\left(x_{0}\right) f\left(y_{0}\right)=0 \tag{3}
\end{align*}
$$

However, by Claim 1 we have $f\left(x_{0}\right) \neq 0$ and $f\left(y_{0}\right) \neq 0$ since $x_{0} \neq 1$ and $y_{0} \neq 1$, a contradiction.

Now the end is near. For any $t \in \mathbb{R}$, plug in $(x, y)=(t,-t)$ in the original equation (*) to get

$$
\begin{aligned}
f(f(t) f(-t))+f(0)=f\left(-t^{2}\right) & \Longleftrightarrow f(f(t) f(-t))=f\left(-t^{2}\right)+1 & & \text { by }(3) \\
& \Longleftrightarrow f(f(t) f(-t))=f\left(-t^{2}+1\right) & & \text { by }(4) \\
& \Longleftrightarrow f(t) f(-t)=-t^{2}+1 & & \text { by injectivity of } f .
\end{aligned}
$$

Similarly, plugging in $(x, y)=(t, 1-t)$ in $(*)$ we get

$$
\begin{aligned}
f(f(t) f(1-t))+f(1)=f(t(1-t)) & \Longleftrightarrow f(f(t) f(1-t))=f(t(1-t)) & \text { by }(3) \\
& \Longleftrightarrow f(t) f(1-t)=t(1-t) \quad & \text { by injectivity of } f .
\end{aligned}
$$

But since $f(1-t)=1+f(-t)$ by (4), we get

$$
\begin{aligned}
f(t) f(1-t)=t(1-t) & \Longleftrightarrow f(t)(1+f(-t))=t(1-t) \Longleftrightarrow f(t)+\left(-t^{2}+1\right)=t(1-t) \\
& \Longleftrightarrow f(t)=t-1,
\end{aligned}
$$

as desired.

Comment. Other approaches are possible. For instance, after Claim 1, we may define

$$
g(x) \stackrel{\text { def }}{=} f(x)+1 .
$$

Replacing $x+1$ and $y+1$ in place of $x$ and $y$ in the original equation (*), we get

$$
f(f(x+1) f(y+1))+f(x+y+2)=f(x y+x+y+1) \quad(x, y \in \mathbb{R})
$$

and therefore, using (4) (so that in particular $g(x)=f(x+1)$ ), we may rewrite ( $*$ ) as

$$
\begin{equation*}
g(g(x) g(y))+g(x+y)=g(x y+x+y) \quad(x, y \in \mathbb{R}) \tag{**}
\end{equation*}
$$

We are now to show that $g(x)=x$ for all $x \in \mathbb{R}$ under the assumption (Claim 1) that 0 is the unique zero of $g$.
Claim 3. Let $n \in \mathbb{Z}$ and $x \in \mathbb{R}$. Then
(a) $g(x+n)=x+n$, and the conditions $g(x)=n$ and $x=n$ are equivalent.
(b) $g(n x)=n g(x)$.

Proof. For part (a), just note that $g(x+n)=x+n$ is just a reformulation of (4). Then $g(x)=n \Longleftrightarrow$ $g(x-n)=0 \Longleftrightarrow x-n=0$ since 0 is the unique zero of $g$. For part (b), we may assume that $x \neq 0$ since the result is obvious when $x=0$. Plug in $y=n / x$ in (**) and use part (a) to get

$$
g\left(g(x) g\left(\frac{n}{x}\right)\right)+g\left(x+\frac{n}{x}\right)=g\left(n+x+\frac{n}{x}\right) \Longleftrightarrow g\left(g(x) g\left(\frac{n}{x}\right)\right)=n \Longleftrightarrow g(x) g\left(\frac{n}{x}\right)=n
$$

In other words, for $x \neq 0$ we have

$$
g(x)=\frac{n}{g(n / x)}
$$

In particular, for $n=1$, we get $g(1 / x)=1 / g(x)$, and therefore replacing $x \leftarrow n x$ in the last equation we finally get

$$
g(n x)=\frac{n}{g(1 / x)}=n g(x)
$$

as required.
Claim 4. The function $g$ is additive, i.e., $g(a+b)=g(a)+g(b)$ for all $a, b \in \mathbb{R}$.
Proof. Set $x \leftarrow-x$ and $y \leftarrow-y$ in $(* *)$; since $g$ is an odd function (by Claim 3(b) with $n=-1$ ), we get

$$
g(g(x) g(y))-g(x+y)=-g(-x y+x+y)
$$

Subtracting the last relation from $(* *)$ we have

$$
2 g(x+y)=g(x y+x+y)+g(-x y+x+y)
$$

and since by Claim 3(b) we have $2 g(x+y)=g(2(x+y))$, we may rewrite the last equation as

$$
g(\alpha+\beta)=g(\alpha)+g(\beta) \quad \text { where } \quad\left\{\begin{array}{l}
\alpha=x y+x+y \\
\beta=-x y+x+y
\end{array}\right.
$$

In other words, we have additivity for all $\alpha, \beta \in \mathbb{R}$ for which there are real numbers $x$ and $y$ satisfying

$$
x+y=\frac{\alpha+\beta}{2} \quad \text { and } \quad x y=\frac{\alpha-\beta}{2}
$$

i.e., for all $\alpha, \beta \in \mathbb{R}$ such that $\left(\frac{\alpha+\beta}{2}\right)^{2}-4 \cdot \frac{\alpha-\beta}{2} \geqslant 0$. Therefore, given any $a, b \in \mathbb{R}$, we may choose $n \in \mathbb{Z}$ large enough so that we have additivity for $\alpha=n a$ and $\beta=n b$, i.e.,

$$
g(n a)+g(n b)=g(n a+n b) \Longleftrightarrow n g(a)+n g(b)=n g(a+b)
$$

by Claim 3(b). Cancelling $n$, we get the desired result. (Alternatively, setting either $(\alpha, \beta)=(a, b)$ or $(\alpha, \beta)=(-a,-b)$ will ensure that $\left.\left(\frac{\alpha+\beta}{2}\right)^{2}-4 \cdot \frac{\alpha-\beta}{2} \geqslant 0\right)$.

Now we may finish the solution. Set $y=1$ in $(* *)$, and use Claim 3 to get

$$
g(g(x) g(1))+g(x+1)=g(2 x+1) \Longleftrightarrow g(g(x))+g(x)+1=2 g(x)+1 \Longleftrightarrow g(g(x))=g(x)
$$

By additivity, this is equivalent to $g(g(x)-x)=0$. Since 0 is the unique zero of $g$ by assumption, we finally get $g(x)-x=0 \Longleftrightarrow g(x)=x$ for all $x \in \mathbb{R}$.

A7. Let $a_{0}, a_{1}, a_{2}, \ldots$ be a sequence of integers and $b_{0}, b_{1}, b_{2}, \ldots$ be a sequence of positive integers such that $a_{0}=0, a_{1}=1$, and

$$
a_{n+1}=\left\{\begin{array}{ll}
a_{n} b_{n}+a_{n-1}, & \text { if } b_{n-1}=1 \\
a_{n} b_{n}-a_{n-1}, & \text { if } b_{n-1}>1
\end{array} \quad \text { for } n=1,2, \ldots\right.
$$

Prove that at least one of the two numbers $a_{2017}$ and $a_{2018}$ must be greater than or equal to 2017 .
(Australia)
Solution 1. The value of $b_{0}$ is irrelevant since $a_{0}=0$, so we may assume that $b_{0}=1$.
Lemma. We have $a_{n} \geqslant 1$ for all $n \geqslant 1$.
Proof. Let us suppose otherwise in order to obtain a contradiction. Let

$$
\begin{equation*}
n \geqslant 1 \text { be the smallest integer with } a_{n} \leqslant 0 \tag{1}
\end{equation*}
$$

Note that $n \geqslant 2$. It follows that $a_{n-1} \geqslant 1$ and $a_{n-2} \geqslant 0$. Thus we cannot have $a_{n}=$ $a_{n-1} b_{n-1}+a_{n-2}$, so we must have $a_{n}=a_{n-1} b_{n-1}-a_{n-2}$. Since $a_{n} \leqslant 0$, we have $a_{n-1} \leqslant a_{n-2}$. Thus we have $a_{n-2} \geqslant a_{n-1} \geqslant a_{n}$.

Let

$$
\begin{equation*}
r \text { be the smallest index with } a_{r} \geqslant a_{r+1} \geqslant a_{r+2} \text {. } \tag{2}
\end{equation*}
$$

Then $r \leqslant n-2$ by the above, but also $r \geqslant 2$ : if $b_{1}=1$, then $a_{2}=a_{1}=1$ and $a_{3}=a_{2} b_{2}+a_{1}>a_{2}$; if $b_{1}>1$, then $a_{2}=b_{1}>1=a_{1}$.

By the minimal choice (2) of $r$, it follows that $a_{r-1}<a_{r}$. And since $2 \leqslant r \leqslant n-2$, by the minimal choice (1) of $n$ we have $a_{r-1}, a_{r}, a_{r+1}>0$. In order to have $a_{r+1} \geqslant a_{r+2}$, we must have $a_{r+2}=a_{r+1} b_{r+1}-a_{r}$ so that $b_{r} \geqslant 2$. Putting everything together, we conclude that

$$
a_{r+1}=a_{r} b_{r} \pm a_{r-1} \geqslant 2 a_{r}-a_{r-1}=a_{r}+\left(a_{r}-a_{r-1}\right)>a_{r},
$$

which contradicts (2).
To complete the problem, we prove that $\max \left\{a_{n}, a_{n+1}\right\} \geqslant n$ by induction. The cases $n=0,1$ are given. Assume it is true for all non-negative integers strictly less than $n$, where $n \geqslant 2$. There are two cases:
Case 1: $b_{n-1}=1$.
Then $a_{n+1}=a_{n} b_{n}+a_{n-1}$. By the inductive assumption one of $a_{n-1}, a_{n}$ is at least $n-1$ and the other, by the lemma, is at least 1 . Hence

$$
a_{n+1}=a_{n} b_{n}+a_{n-1} \geqslant a_{n}+a_{n-1} \geqslant(n-1)+1=n .
$$

Thus $\max \left\{a_{n}, a_{n+1}\right\} \geqslant n$, as desired.
Case 2: $b_{n-1}>1$.
Since we defined $b_{0}=1$ there is an index $r$ with $1 \leqslant r \leqslant n-1$ such that

$$
b_{n-1}, b_{n-2}, \ldots, b_{r} \geqslant 2 \quad \text { and } \quad b_{r-1}=1
$$

We have $a_{r+1}=a_{r} b_{r}+a_{r-1} \geqslant 2 a_{r}+a_{r-1}$. Thus $a_{r+1}-a_{r} \geqslant a_{r}+a_{r-1}$.
Now we claim that $a_{r}+a_{r-1} \geqslant r$. Indeed, this holds by inspection for $r=1$; for $r \geqslant 2$, one of $a_{r}, a_{r-1}$ is at least $r-1$ by the inductive assumption, while the other, by the lemma, is at least 1 . Hence $a_{r}+a_{r-1} \geqslant r$, as claimed, and therefore $a_{r+1}-a_{r} \geqslant r$ by the last inequality in the previous paragraph.

Since $r \geqslant 1$ and, by the lemma, $a_{r} \geqslant 1$, from $a_{r+1}-a_{r} \geqslant r$ we get the following two inequalities:

$$
a_{r+1} \geqslant r+1 \quad \text { and } \quad a_{r+1}>a_{r} .
$$

Now observe that

$$
a_{m}>a_{m-1} \Longrightarrow a_{m+1}>a_{m} \text { for } m=r+1, r+2, \ldots, n-1,
$$

since $a_{m+1}=a_{m} b_{m}-a_{m-1} \geqslant 2 a_{m}-a_{m-1}=a_{m}+\left(a_{m}-a_{m-1}\right)>a_{m}$. Thus

$$
a_{n}>a_{n-1}>\cdots>a_{r+1} \geqslant r+1 \Longrightarrow a_{n} \geqslant n .
$$

So $\max \left\{a_{n}, a_{n+1}\right\} \geqslant n$, as desired.
Solution 2. We say that an index $n>1$ is bad if $b_{n-1}=1$ and $b_{n-2}>1$; otherwise $n$ is good. The value of $b_{0}$ is irrelevant to the definition of $\left(a_{n}\right)$ since $a_{0}=0$; so we assume that $b_{0}>1$.
Lemma 1. (a) $a_{n} \geqslant 1$ for all $n>0$.
(b) If $n>1$ is good, then $a_{n}>a_{n-1}$.

Proof. Induction on $n$. In the base cases $n=1,2$ we have $a_{1}=1 \geqslant 1, a_{2}=b_{1} a_{1} \geqslant 1$, and finally $a_{2}>a_{1}$ if 2 is good, since in this case $b_{1}>1$.

Now we assume that the lemma statement is proved for $n=1,2, \ldots, k$ with $k \geqslant 2$, and prove it for $n=k+1$. Recall that $a_{k}$ and $a_{k-1}$ are positive by the induction hypothesis.
Case 1: $k$ is bad.
We have $b_{k-1}=1$, so $a_{k+1}=b_{k} a_{k}+a_{k-1} \geqslant a_{k}+a_{k-1}>a_{k} \geqslant 1$, as required.
Case 2: $k$ is good.
We already have $a_{k}>a_{k-1} \geqslant 1$ by the induction hypothesis. We consider three easy subcases.

Subcase 2.1: $b_{k}>1$.
Then $a_{k+1} \geqslant b_{k} a_{k}-a_{k-1} \geqslant a_{k}+\left(a_{k}-a_{k-1}\right)>a_{k} \geqslant 1$.
Subcase 2.2: $b_{k}=b_{k-1}=1$.
Then $a_{k+1}=a_{k}+a_{k-1}>a_{k} \geqslant 1$.
Subcase 2.3: $b_{k}=1$ but $b_{k-1}>1$.
Then $k+1$ is bad, and we need to prove only (a), which is trivial: $a_{k+1}=a_{k}-a_{k-1} \geqslant 1$.
So, in all three subcases we have verified the required relations.
Lemma 2. Assume that $n>1$ is bad. Then there exists a $j \in\{1,2,3\}$ such that $a_{n+j} \geqslant$ $a_{n-1}+j+1$, and $a_{n+i} \geqslant a_{n-1}+i$ for all $1 \leqslant i<j$.
Proof. Recall that $b_{n-1}=1$. Set

$$
m=\inf \left\{i>0: b_{n+i-1}>1\right\}
$$

(possibly $m=+\infty$ ). We claim that $j=\min \{m, 3\}$ works. Again, we distinguish several cases, according to the value of $m$; in each of them we use Lemma 1 without reference.
Case 1: $m=1$, so $b_{n}>1$.
Then $a_{n+1} \geqslant 2 a_{n}+a_{n-1} \geqslant a_{n-1}+2$, as required.
Case 2: $m=2$, so $b_{n}=1$ and $b_{n+1}>1$.
Then we successively get

$$
\begin{gathered}
a_{n+1}=a_{n}+a_{n-1} \geqslant a_{n-1}+1 \\
a_{n+2} \geqslant 2 a_{n+1}+a_{n} \geqslant 2\left(a_{n-1}+1\right)+a_{n}=a_{n-1}+\left(a_{n-1}+a_{n}+2\right) \geqslant a_{n-1}+4,
\end{gathered}
$$

which is even better than we need.

Case 3: $m>2$, so $b_{n}=b_{n+1}=1$.
Then we successively get

$$
\begin{gathered}
a_{n+1}=a_{n}+a_{n-1} \geqslant a_{n-1}+1, \quad a_{n+2}=a_{n+1}+a_{n} \geqslant a_{n-1}+1+a_{n} \geqslant a_{n-1}+2, \\
a_{n+3} \geqslant a_{n+2}+a_{n+1} \geqslant\left(a_{n-1}+1\right)+\left(a_{n-1}+2\right) \geqslant a_{n-1}+4,
\end{gathered}
$$

as required.
Lemmas 1 (b) and 2 provide enough information to prove that $\max \left\{a_{n}, a_{n+1}\right\} \geqslant n$ for all $n$ and, moreover, that $a_{n} \geqslant n$ often enough. Indeed, assume that we have found some $n$ with $a_{n-1} \geqslant n-1$. If $n$ is good, then by Lemma $1(\mathrm{~b})$ we have $a_{n} \geqslant n$ as well. If $n$ is bad, then Lemma 2 yields $\max \left\{a_{n+i}, a_{n+i+1}\right\} \geqslant a_{n-1}+i+1 \geqslant n+i$ for all $0 \leqslant i<j$ and $a_{n+j} \geqslant a_{n-1}+j+1 \geqslant n+j$; so $n+j$ is the next index to start with.

A8. Assume that a function $f: \mathbb{R} \rightarrow \mathbb{R}$ satisfies the following condition:
For every $x, y \in \mathbb{R}$ such that $(f(x)+y)(f(y)+x)>0$, we have $f(x)+y=f(y)+x$.
Prove that $f(x)+y \leqslant f(y)+x$ whenever $x>y$.
(Netherlands)
Solution 1. Define $g(x)=x-f(x)$. The condition on $f$ then rewrites as follows:
For every $x, y \in \mathbb{R}$ such that $((x+y)-g(x))((x+y)-g(y))>0$, we have $g(x)=g(y)$.
This condition may in turn be rewritten in the following form:
If $g(x) \neq g(y)$, then the number $x+y$ lies (non-strictly) between $g(x)$ and $g(y)$.
Notice here that the function $g_{1}(x)=-g(-x)$ also satisfies $(*)$, since

$$
\begin{aligned}
g_{1}(x) \neq g_{1}(y) \Longrightarrow \quad & g(-x) \neq g(-y) \Longrightarrow \quad \Longrightarrow \quad(x+y) \text { lies between } g(-x) \text { and } g(-y) \\
& \Longrightarrow x+y \text { lies between } g_{1}(x) \text { and } g_{1}(y) .
\end{aligned}
$$

On the other hand, the relation we need to prove reads now as

$$
\begin{equation*}
g(x) \leqslant g(y) \quad \text { whenever } x<y \tag{1}
\end{equation*}
$$

Again, this condition is equivalent to the same one with $g$ replaced by $g_{1}$.
If $g(x)=2 x$ for all $x \in \mathbb{R}$, then $(*)$ is obvious; so in what follows we consider the other case. We split the solution into a sequence of lemmas, strengthening one another. We always consider some value of $x$ with $g(x) \neq 2 x$ and denote $X=g(x)$.
Lemma 1. Assume that $X<2 x$. Then on the interval $(X-x ; x]$ the function $g$ attains at most two values - namely, $X$ and, possibly, some $Y>X$. Similarly, if $X>2 x$, then $g$ attains at most two values on $[x ; X-x)$ - namely, $X$ and, possibly, some $Y<X$.
Proof. We start with the first claim of the lemma. Notice that $X-x<x$, so the considered interval is nonempty.

Take any $a \in(X-x ; x)$ with $g(a) \neq X$ (if it exists). If $g(a)<X$, then (*) yields $g(a) \leqslant$ $a+x \leqslant g(x)=X$, so $a \leqslant X-x$ which is impossible. Thus, $g(a)>X$ and hence by (*) we get $X \leqslant a+x \leqslant g(a)$.

Now, for any $b \in(X-x ; x)$ with $g(b) \neq X$ we similarly get $b+x \leqslant g(b)$. Therefore, the number $a+b$ (which is smaller than each of $a+x$ and $b+x$ ) cannot lie between $g(a)$ and $g(b)$, which by $(*)$ implies that $g(a)=g(b)$. Hence $g$ may attain only two values on $(X-x ; x]$, namely $X$ and $g(a)>X$.

To prove the second claim, notice that $g_{1}(-x)=-X<2 \cdot(-x)$, so $g_{1}$ attains at most two values on $(-X+x,-x]$, i.e., $-X$ and, possibly, some $-Y>-X$. Passing back to $g$, we get what we need.
Lemma 2. If $X<2 x$, then $g$ is constant on $(X-x ; x)$. Similarly, if $X>2 x$, then $g$ is constant on $(x ; X-x)$.
Proof. Again, it suffices to prove the first claim only. Assume, for the sake of contradiction, that there exist $a, b \in(X-x ; x)$ with $g(a) \neq g(b)$; by Lemma 1, we may assume that $g(a)=X$ and $Y=g(b)>X$.

Notice that $\min \{X-a, X-b\}>X-x$, so there exists a $u \in(X-x ; x)$ such that $u<\min \{X-a, X-b\}$. By Lemma 1, we have either $g(u)=X$ or $g(u)=Y$. In the former case, by (*) we have $X \leqslant u+b \leqslant Y$ which contradicts $u<X-b$. In the second case, by (*) we have $X \leqslant u+a \leqslant Y$ which contradicts $u<X-a$. Thus the lemma is proved.

Lemma 3. If $X<2 x$, then $g(a)=X$ for all $a \in(X-x ; x)$. Similarly, if $X>2 x$, then $g(a)=X$ for all $a \in(x ; X-x)$.
Proof. Again, we only prove the first claim.
By Lemmas 1 and 2, this claim may be violated only if $g$ takes on a constant value $Y>X$ on ( $X-x, x$ ). Choose any $a, b \in(X-x ; x)$ with $a<b$. By (*), we have

$$
\begin{equation*}
Y \geqslant b+x \geqslant X \tag{2}
\end{equation*}
$$

In particular, we have $Y \geqslant b+x>2 a$. Applying Lemma 2 to $a$ in place of $x$, we obtain that $g$ is constant on $(a, Y-a)$. By (2) again, we have $x \leqslant Y-b<Y-a$; so $x, b \in(a ; Y-a)$. But $X=g(x) \neq g(b)=Y$, which is a contradiction.

Now we are able to finish the solution. Assume that $g(x)>g(y)$ for some $x<y$. Denote $X=g(x)$ and $Y=g(y)$; by (*), we have $X \geqslant x+y \geqslant Y$, so $Y-y \leqslant x<y \leqslant X-x$, and hence $(Y-y ; y) \cap(x ; X-x)=(x, y) \neq \varnothing$. On the other hand, since $Y-y<y$ and $x<X-x$, Lemma 3 shows that $g$ should attain a constant value $X$ on $(x ; X-x)$ and a constant value $Y \neq X$ on $(Y-y ; y)$. Since these intervals overlap, we get the final contradiction.

Solution 2. As in the previous solution, we pass to the function $g$ satisfying (*) and notice that we need to prove the condition (1). We will also make use of the function $g_{1}$.

If $g$ is constant, then (1) is clearly satisfied. So, in the sequel we assume that $g$ takes on at least two different values. Now we collect some information about the function $g$.
Claim 1. For any $c \in \mathbb{R}$, all the solutions of $g(x)=c$ are bounded.
Proof. Fix any $y \in \mathbb{R}$ with $g(y) \neq c$. Assume first that $g(y)>c$. Now, for any $x$ with $g(x)=c$, by (*) we have $c \leqslant x+y \leqslant g(y)$, or $c-y \leqslant x \leqslant g(y)-y$. Since $c$ and $y$ are constant, we get what we need.

If $g(y)<c$, we may switch to the function $g_{1}$ for which we have $g_{1}(-y)>-c$. By the above arguments, we obtain that all the solutions of $g_{1}(-x)=-c$ are bounded, which is equivalent to what we need.

As an immediate consequence, the function $g$ takes on infinitely many values, which shows that the next claim is indeed widely applicable.
Claim 2. If $g(x)<g(y)<g(z)$, then $x<z$.
Proof. By (*), we have $g(x) \leqslant x+y \leqslant g(y) \leqslant z+y \leqslant g(z)$, so $x+y \leqslant z+y$, as required.
Claim 3. Assume that $g(x)>g(y)$ for some $x<y$. Then $g(a) \in\{g(x), g(y)\}$ for all $a \in[x ; y]$.
Proof. If $g(y)<g(a)<g(x)$, then the triple ( $y, a, x)$ violates Claim 2. If $g(a)<g(y)<g(x)$, then the triple $(a, y, x)$ violates Claim 2. If $g(y)<g(x)<g(a)$, then the triple ( $y, x, a$ ) violates Claim 2. The only possible cases left are $g(a) \in\{g(x), g(y)\}$.

In view of Claim 3, we say that an interval $I$ (which may be open, closed, or semi-open) is a Dirichlet interval ${ }^{*}$ if the function $g$ takes on just two values on $I$.

Assume now, for the sake of contradiction, that (1) is violated by some $x<y$. By Claim 3, $[x ; y]$ is a Dirichlet interval. Set
$r=\inf \{a:(a ; y]$ is a Dirichlet interval $\}$ and $s=\sup \{b:[x ; b)$ is a Dirichlet interval $\}$.
Clearly, $r \leqslant x<y \leqslant s$. By Claim 1, $r$ and $s$ are finite. Denote $X=g(x), Y=g(y)$, and $\Delta=(y-x) / 2$.

Suppose first that there exists a $t \in(r ; r+\Delta)$ with $f(t)=Y$. By the definition of $r$, the interval $(r-\Delta ; y]$ is not Dirichlet, so there exists an $r^{\prime} \in(r-\Delta ; r]$ such that $g\left(r^{\prime}\right) \notin\{X, Y\}$.

[^0]The function $g$ attains at least three distinct values on $\left[r^{\prime} ; y\right]$, namely $g\left(r^{\prime}\right), g(x)$, and $g(y)$. Claim 3 now yields $g\left(r^{\prime}\right) \leqslant g(y)$; the equality is impossible by the choice of $r^{\prime}$, so in fact $g\left(r^{\prime}\right)<Y$. Applying (*) to the pairs $\left(r^{\prime}, y\right)$ and $(t, x)$ we obtain $r^{\prime}+y \leqslant Y \leqslant t+x$, whence $r-\Delta+y<r^{\prime}+y \leqslant t+x<r+\Delta+x$, or $y-x<2 \Delta$. This is a contradiction.

Thus, $g(t)=X$ for all $t \in(r ; r+\Delta)$. Applying the same argument to $g_{1}$, we get $g(t)=Y$ for all $t \in(s-\Delta ; s)$.

Finally, choose some $s_{1}, s_{2} \in(s-\Delta ; s)$ with $s_{1}<s_{2}$ and denote $\delta=\left(s_{2}-s_{1}\right) / 2$. As before, we choose $r^{\prime} \in(r-\delta ; r)$ with $g\left(r^{\prime}\right) \notin\{X, Y\}$ and obtain $g\left(r^{\prime}\right)<Y$. Choose any $t \in(r ; r+\delta)$; by the above arguments, we have $g(t)=X$ and $g\left(s_{1}\right)=g\left(s_{2}\right)=Y$. As before, we apply (*) to the pairs $\left(r^{\prime}, s_{2}\right)$ and $\left(t, s_{1}\right)$ obtaining $r-\delta+s_{2}<r^{\prime}+s_{2} \leqslant Y \leqslant t+s_{1}<r+\delta+s_{1}$, or $s_{2}-s_{1}<2 \delta$. This is a final contradiction.

Comment 1. The original submission discussed the same functions $f$, but the question was different - namely, the following one:

Prove that the equation $f(x)=2017 x$ has at most one solution, and the equation $f(x)=-2017 x$ has at least one solution.

The Problem Selection Committee decided that the question we are proposing is more natural, since it provides more natural information about the function $g$ (which is indeed the main character in this story). On the other hand, the new problem statement is strong enough in order to imply the original one easily.

Namely, we will deduce from the new problem statement (along with the facts used in the solutions) that ( $i$ ) for every $N>0$ the equation $g(x)=-N x$ has at most one solution, and (ii) for every $N>1$ the equation $g(x)=N x$ has at least one solution.

Claim ( $i$ ) is now trivial. Indeed, $g$ is proven to be non-decreasing, so $g(x)+N x$ is strictly increasing and thus has at most one zero.

We proceed on claim $(i i)$. If $g(0)=0$, then the required root has been already found. Otherwise, we may assume that $g(0)>0$ and denote $c=g(0)$. We intend to prove that $x=c / N$ is the required root. Indeed, by monotonicity we have $g(c / N) \geqslant g(0)=c$; if we had $g(c / N)>c$, then (*) would yield $c \leqslant 0+c / N \leqslant g(c / N)$ which is false. Thus, $g(x)=c=N x$.

Comment 2. There are plenty of functions $g$ satisfying (*) (and hence of functions $f$ satisfying the problem conditions). One simple example is $g_{0}(x)=2 x$. Next, for any increasing sequence $A=\left(\ldots, a_{-1}, a_{0}, a_{1}, \ldots\right)$ which is unbounded in both directions (i.e., for every $N$ this sequence contains terms greater than $N$, as well as terms smaller than $-N$ ), the function $g_{A}$ defined by

$$
g_{A}(x)=a_{i}+a_{i+1} \quad \text { whenever } x \in\left[a_{i} ; a_{i+1}\right)
$$

satisfies (*). Indeed, pick any $x<y$ with $g(x) \neq g(y)$; this means that $x \in\left[a_{i} ; a_{i+1}\right)$ and $y \in\left[a_{j} ; a_{j+1}\right)$ for some $i<j$. Then we have $g(x)=a_{i}+a_{i+1} \leqslant x+y<a_{j}+a_{j+1}=g(y)$, as required.

There also exist examples of the mixed behavior; e.g., for an arbitrary sequence $A$ as above and an arbitrary subset $I \subseteq \mathbb{Z}$ the function

$$
g_{A, I}(x)=\left\{\begin{array}{lll}
g_{0}(x), & x \in\left[a_{i} ; a_{i+1}\right) & \text { with } i \in I ; \\
g_{A}(x), & x \in\left[a_{i} ; a_{i+1}\right) & \text { with } i \notin I
\end{array}\right.
$$

also satisfies (*).
Finally, it is even possible to provide a complete description of all functions $g$ satisfying (*) (and hence of all functions $f$ satisfying the problem conditions); however, it seems to be far out of scope for the IMO. This description looks as follows.

Let $A$ be any closed subset of $\mathbb{R}$ which is unbounded in both directions. Define the functions $i_{A}$, $s_{A}$, and $g_{A}$ as follows:

$$
i_{A}(x)=\inf \{a \in A: a \geqslant x\}, \quad s_{A}(x)=\sup \{a \in A: a \leqslant x\}, \quad g_{A}(x)=i_{A}(x)+s_{A}(x) .
$$

It is easy to see that for different sets $A$ and $B$ the functions $g_{A}$ and $g_{B}$ are also different (since, e.g., for any $a \in A \backslash B$ the function $g_{B}$ is constant in a small neighborhood of $a$, but the function $g_{A}$ is not). One may check, similarly to the arguments above, that each such function satisfies (*).

Finally, one more modification is possible. Namely, for any $x \in A$ one may redefine $g_{A}(x)$ (which is $2 x$ ) to be any of the numbers

$$
\begin{gathered}
\quad g_{A+}(x)=i_{A+}(x)+x \quad \text { or } \quad g_{A-}(x)=x+s_{A-}(x), \\
\text { where } \quad i_{A+}(x)=\inf \{a \in A: a>x\} \quad \text { and } \quad s_{A-}(x)=\sup \{a \in A: a<x\} .
\end{gathered}
$$

This really changes the value if $x$ has some right (respectively, left) semi-neighborhood disjoint from $A$, so there are at most countably many possible changes; all of them can be performed independently.

With some effort, one may show that the construction above provides all functions $g$ satisfying (*).

## Combinatorics

C1. A rectangle $\mathcal{R}$ with odd integer side lengths is divided into small rectangles with integer side lengths. Prove that there is at least one among the small rectangles whose distances from the four sides of $\mathcal{R}$ are either all odd or all even.
(Singapore)
Solution. Let the width and height of $\mathcal{R}$ be odd numbers $a$ and $b$. Divide $\mathcal{R}$ into $a b$ unit squares and color them green and yellow in a checkered pattern. Since the side lengths of $a$ and $b$ are odd, the corner squares of $\mathcal{R}$ will all have the same color, say green.

Call a rectangle (either $\mathcal{R}$ or a small rectangle) green if its corners are all green; call it yellow if the corners are all yellow, and call it mixed if it has both green and yellow corners. In particular, $\mathcal{R}$ is a green rectangle.

We will use the following trivial observations.

- Every mixed rectangle contains the same number of green and yellow squares;
- Every green rectangle contains one more green square than yellow square;
- Every yellow rectangle contains one more yellow square than green square.

The rectangle $\mathcal{R}$ is green, so it contains more green unit squares than yellow unit squares. Therefore, among the small rectangles, at least one is green. Let $\mathcal{S}$ be such a small green rectangle, and let its distances from the sides of $\mathcal{R}$ be $x, y, u$ and $v$, as shown in the picture. The top-left corner of $\mathcal{R}$ and the top-left corner of $\mathcal{S}$ have the same color, which happen if and only if $x$ and $u$ have the same parity. Similarly, the other three green corners of $\mathcal{S}$ indicate that $x$ and $v$ have the same parity, $y$ and $u$ have the same parity, i.e. $x, y, u$ and $v$ are all odd or all even.


C2. Let $n$ be a positive integer. Define a chameleon to be any sequence of $3 n$ letters, with exactly $n$ occurrences of each of the letters $a, b$, and $c$. Define a swap to be the transposition of two adjacent letters in a chameleon. Prove that for any chameleon $X$, there exists a chameleon $Y$ such that $X$ cannot be changed to $Y$ using fewer than $3 n^{2} / 2$ swaps.
(Australia)
Solution 1. To start, notice that the swap of two identical letters does not change a chameleon, so we may assume there are no such swaps.

For any two chameleons $X$ and $Y$, define their distance $d(X, Y)$ to be the minimal number of swaps needed to transform $X$ into $Y$ (or vice versa). Clearly, $d(X, Y)+d(Y, Z) \geqslant d(X, Z)$ for any three chameleons $X, Y$, and $Z$.
Lemma. Consider two chameleons

$$
P=\underbrace{a a \ldots a}_{n} \underbrace{b b \ldots b}_{n} \underbrace{c c \ldots c}_{n} \text { and } Q=\underbrace{c c \ldots c}_{n} \underbrace{b b \ldots b}_{n} \underbrace{a a \ldots a}_{n} .
$$

Then $d(P, Q) \geqslant 3 n^{2}$.
Proof. For any chameleon $X$ and any pair of distinct letters $u, v \in\{a, b, c\}$, we define $f_{u, v}(X)$ to be the number of pairs of positions in $X$ such that the left one is occupied by $u$, and the right one is occupied by $v$. Define $f(X)=f_{a, b}(X)+f_{a, c}(X)+f_{b, c}(X)$. Notice that $f_{a, b}(P)=f_{a, c}(P)=f_{b, c}(P)=n^{2}$ and $f_{a, b}(Q)=f_{a, c}(Q)=f_{b, c}(Q)=0$, so $f(P)=3 n^{2}$ and $f(Q)=0$.

Now consider some swap changing a chameleon $X$ to $X^{\prime}$; say, the letters $a$ and $b$ are swapped. Then $f_{a, b}(X)$ and $f_{a, b}\left(X^{\prime}\right)$ differ by exactly 1 , while $f_{a, c}(X)=f_{a, c}\left(X^{\prime}\right)$ and $f_{b, c}(X)=f_{b, c}\left(X^{\prime}\right)$. This yields $\left|f(X)-f\left(X^{\prime}\right)\right|=1$, i.e., on any swap the value of $f$ changes by 1 . Hence $d(X, Y) \geqslant$ $|f(X)-f(Y)|$ for any two chameleons $X$ and $Y$. In particular, $d(P, Q) \geqslant|f(P)-f(Q)|=3 n^{2}$, as desired.

Back to the problem, take any chameleon $X$ and notice that $d(X, P)+d(X, Q) \geqslant d(P, Q) \geqslant$ $3 n^{2}$ by the lemma. Consequently, $\max \{d(X, P), d(X, Q)\} \geqslant \frac{3 n^{2}}{2}$, which establishes the problem statement.

Comment 1. The problem may be reformulated in a graph language. Construct a graph $G$ with the chameleons as vertices, two vertices being connected with an edge if and only if these chameleons differ by a single swap. Then $d(X, Y)$ is the usual distance between the vertices $X$ and $Y$ in this graph. Recall that the radius of a connected graph $G$ is defined as

$$
r(G)=\min _{v \in V} \max _{u \in V} d(u, v) .
$$

So we need to prove that the radius of the constructed graph is at least $3 n^{2} / 2$.
It is well-known that the radius of any connected graph is at least the half of its diameter (which is simply $\max _{u, v \in V} d(u, v)$ ). Exactly this fact has been used above in order to finish the solution.

Solution 2. We use the notion of distance from Solution 1, but provide a different lower bound for it.

In any chameleon $X$, we enumerate the positions in it from left to right by $1,2, \ldots, 3 n$. Define $s_{c}(X)$ as the sum of positions occupied by $c$. The value of $s_{c}$ changes by at most 1 on each swap, but this fact alone does not suffice to solve the problem; so we need an improvement.

For every chameleon $X$, denote by $X_{\bar{c}}$ the sequence obtained from $X$ by removing all $n$ letters $c$. Enumerate the positions in $X_{\bar{c}}$ from left to right by $1,2, \ldots, 2 n$, and define $s_{\bar{c}, b}(X)$ as the sum of positions in $X_{\bar{c}}$ occupied by $b$. (In other words, here we consider the positions of the $b$ 's relatively to the $a$ 's only.) Finally, denote

$$
d^{\prime}(X, Y):=\left|s_{c}(X)-s_{c}(Y)\right|+\left|s_{\bar{c}, b}(X)-s_{\bar{c}, b}(Y)\right| .
$$

Now consider any swap changing a chameleon $X$ to $X^{\prime}$. If no letter $c$ is involved into this swap, then $s_{c}(X)=s_{c}\left(X^{\prime}\right)$; on the other hand, exactly one letter $b$ changes its position in $X_{\bar{c}}$, so $\left|s_{\bar{c}, b}(X)-s_{\bar{c}, b}\left(X^{\prime}\right)\right|=1$. If a letter $c$ is involved into a swap, then $X_{\bar{c}}=X_{\bar{c}}^{\prime}$, so $s_{\bar{c}, b}(X)=s_{\bar{c}, b}\left(X^{\prime}\right)$ and $\left|s_{c}(X)-s_{c}\left(X^{\prime}\right)\right|=1$. Thus, in all cases we have $d^{\prime}\left(X, X^{\prime}\right)=1$.

As in the previous solution, this means that $d(X, Y) \geqslant d^{\prime}(X, Y)$ for any two chameleons $X$ and $Y$. Now, for any chameleon $X$ we will indicate a chameleon $Y$ with $d^{\prime}(X, Y) \geqslant 3 n^{2} / 2$, thus finishing the solution.

The function $s_{c}$ attains all integer values from $1+\cdots+n=\frac{n(n+1)}{2}$ to $(2 n+1)+\cdots+3 n=$ $2 n^{2}+\frac{n(n+1)}{2}$. If $s_{c}(X) \leqslant n^{2}+\frac{n(n+1)}{2}$, then we put the letter $c$ into the last $n$ positions in $Y$; otherwise we put the letter $c$ into the first $n$ positions in $Y$. In either case we already have $\left|s_{c}(X)-s_{c}(Y)\right| \geqslant n^{2}$.

Similarly, $s_{\bar{c}, b}$ ranges from $\frac{n(n+1)}{2}$ to $n^{2}+\frac{n(n+1)}{2}$. So, if $s_{\bar{c}, b}(X) \leqslant \frac{n^{2}}{2}+\frac{n(n+1)}{2}$, then we put the letter $b$ into the last $n$ positions in $Y$ which are still free; otherwise, we put the letter $b$ into the first $n$ such positions. The remaining positions are occupied by $a$. In any case, we have $\left|s_{\bar{c}, b}(X)-s_{\bar{c}, b}(Y)\right| \geqslant \frac{n^{2}}{2}$, thus $d^{\prime}(X, Y) \geqslant n^{2}+\frac{n^{2}}{2}=\frac{3 n^{2}}{2}$, as desired.

Comment 2. The two solutions above used two lower bounds $|f(X)-f(Y)|$ and $d^{\prime}(X, Y)$ for the number $d(X, Y)$. One may see that these bounds are closely related to each other, as

$$
f_{a, c}(X)+f_{b, c}(X)=s_{c}(X)-\frac{n(n+1)}{2} \quad \text { and } \quad f_{a, b}(X)=s_{\bar{c}, b}(X)-\frac{n(n+1)}{2} .
$$

One can see that, e.g., the bound $d^{\prime}(X, Y)$ could as well be used in the proof of the lemma in Solution 1.
Let us describe here an even sharper bound which also can be used in different versions of the solutions above.

In each chameleon $X$, enumerate the occurrences of $a$ from the left to the right as $a_{1}, a_{2}, \ldots, a_{n}$. Since we got rid of swaps of identical letters, the relative order of these letters remains the same during the swaps. Perform the same operation with the other letters, obtaining new letters $b_{1}, \ldots, b_{n}$ and $c_{1}, \ldots, c_{n}$. Denote by $A$ the set of the $3 n$ obtained letters.

Since all $3 n$ letters became different, for any chameleon $X$ and any $s \in A$ we may define the position $N_{s}(X)$ of $s$ in $X$ (thus $1 \leqslant N_{s}(X) \leqslant 3 n$ ). Now, for any two chameleons $X$ and $Y$ we say that a pair of letters $(s, t) \in A \times A$ is an $(X, Y)$-inversion if $N_{s}(X)<N_{t}(X)$ but $N_{s}(Y)>N_{t}(Y)$, and define $d^{*}(X, Y)$ to be the number of $(X, Y)$-inversions. Then for any two chameleons $Y$ and $Y^{\prime}$ differing by a single swap, we have $\left|d^{*}(X, Y)-d^{*}\left(X, Y^{\prime}\right)\right|=1$. Since $d^{*}(X, X)=0$, this yields $d(X, Y) \geqslant d^{*}(X, Y)$ for any pair of chameleons $X$ and $Y$. The bound $d^{*}$ may also be used in both Solution 1 and Solution 2.

Comment 3. In fact, one may prove that the distance $d^{*}$ defined in the previous comment coincides with $d$. Indeed, if $X \neq Y$, then there exist an ( $X, Y$ )-inversion $(s, t)$. One can show that such $s$ and $t$ may be chosen to occupy consecutive positions in $Y$. Clearly, $s$ and $t$ correspond to different letters among $\{a, b, c\}$. So, swapping them in $Y$ we get another chameleon $Y^{\prime}$ with $d^{*}\left(X, Y^{\prime}\right)=d^{*}(X, Y)-1$. Proceeding in this manner, we may change $Y$ to $X$ in $d^{*}(X, Y)$ steps.

Using this fact, one can show that the estimate in the problem statement is sharp for all $n \geqslant 2$. (For $n=1$ it is not sharp, since any permutation of three letters can be changed to an opposite one in no less than three swaps.) We outline the proof below.

For any $k \geqslant 0$, define

$$
X_{2 k}=\underbrace{a b c a b c \ldots a b c}_{3 k \text { letters }} \underbrace{c b a c b a \ldots c b a}_{3 k \text { letters }} \quad \text { and } \quad X_{2 k+3}=\underbrace{a b c a b c \ldots a b c}_{3 k \text { letters }} a b c b c a c a b \underbrace{c b a c b a \ldots c b a}_{3 k \text { letters }} .
$$

We claim that for every $n \geqslant 2$ and every chameleon $Y$, we have $d^{*}\left(X_{n}, Y\right) \leqslant\left\lceil 3 n^{2} / 2\right\rceil$. This will mean that for every $n \geqslant 2$ the number $3 n^{2} / 2$ in the problem statement cannot be changed by any number larger than $\left\lceil 3 n^{2} / 2\right\rceil$.

For any distinct letters $u, v \in\{a, b, c\}$ and any two chameleons $X$ and $Y$, we define $d_{u, v}^{*}(X, Y)$ as the number of $(X, Y)$-inversions $(s, t)$ such that $s$ and $t$ are instances of $u$ and $v$ (in any of the two possible orders). Then $d^{*}(X, Y)=d_{a, b}^{*}(X, Y)+d_{b, c}^{*}(X, Y)+d_{c, a}^{*}(X, Y)$.

We start with the case when $n=2 k$ is even; denote $X=X_{2 k}$. We show that $d_{a, b}^{*}(X, Y) \leqslant 2 k^{2}$ for any chameleon $Y$; this yields the required estimate. Proceed by the induction on $k$ with the trivial base case $k=0$. To perform the induction step, notice that $d_{a, b}^{*}(X, Y)$ is indeed the minimal number of swaps needed to change $Y_{\bar{c}}$ into $X_{\bar{c}}$. One may show that moving $a_{1}$ and $a_{2 k}$ in $Y$ onto the first and the last positions in $Y$, respectively, takes at most $2 k$ swaps, and that subsequent moving $b_{1}$ and $b_{2 k}$ onto the second and the second last positions takes at most $2 k-2$ swaps. After performing that, one may delete these letters from both $X_{\bar{c}}$ and $Y_{\bar{c}}$ and apply the induction hypothesis; so $X_{\bar{c}}$ can be obtained from $Y_{\bar{c}}$ using at most $2(k-1)^{2}+2 k+(2 k-2)=2 k^{2}$ swaps, as required.

If $n=2 k+3$ is odd, the proof is similar but more technically involved. Namely, we claim that $d_{a, b}^{*}\left(X_{2 k+3}, Y\right) \leqslant 2 k^{2}+6 k+5$ for any chameleon $Y$, and that the equality is achieved only if $Y_{\bar{c}}=$ $b b \ldots b a a \ldots a$. The proof proceeds by a similar induction, with some care taken of the base case, as well as of extracting the equality case. Similar estimates hold for $d_{b, c}^{*}$ and $d_{c, a}^{*}$. Summing three such estimates, we obtain

$$
d^{*}\left(X_{2 k+3}, Y\right) \leqslant 3\left(2 k^{2}+6 k+5\right)=\left\lceil\frac{3 n^{2}}{2}\right\rceil+1,
$$

which is by 1 more than we need. But the equality could be achieved only if $Y_{\bar{c}}=b b \ldots b a a \ldots a$ and, similarly, $Y_{\bar{b}}=a a \ldots a c c \ldots c$ and $Y_{\bar{a}}=c c \ldots c b b \ldots b$. Since these three equalities cannot hold simultaneously, the proof is finished.

C3. Sir Alex plays the following game on a row of 9 cells. Initially, all cells are empty. In each move, Sir Alex is allowed to perform exactly one of the following two operations:
(1) Choose any number of the form $2^{j}$, where $j$ is a non-negative integer, and put it into an empty cell.
(2) Choose two (not necessarily adjacent) cells with the same number in them; denote that number by $2^{j}$. Replace the number in one of the cells with $2^{j+1}$ and erase the number in the other cell.

At the end of the game, one cell contains the number $2^{n}$, where $n$ is a given positive integer, while the other cells are empty. Determine the maximum number of moves that Sir Alex could have made, in terms of $n$.
(Thailand)
Answer: $2 \sum_{j=0}^{8}\binom{n}{j}-1$.
Solution 1. We will solve a more general problem, replacing the row of 9 cells with a row of $k$ cells, where $k$ is a positive integer. Denote by $m(n, k)$ the maximum possible number of moves Sir Alex can make starting with a row of $k$ empty cells, and ending with one cell containing the number $2^{n}$ and all the other $k-1$ cells empty. Call an operation of type (1) an insertion, and an operation of type (2) a merge.

Only one move is possible when $k=1$, so we have $m(n, 1)=1$. From now on we consider $k \geqslant 2$, and we may assume Sir Alex's last move was a merge. Then, just before the last move, there were exactly two cells with the number $2^{n-1}$, and the other $k-2$ cells were empty.

Paint one of those numbers $2^{n-1}$ blue, and the other one red. Now trace back Sir Alex's moves, always painting the numbers blue or red following this rule: if $a$ and $b$ merge into $c$, paint $a$ and $b$ with the same color as $c$. Notice that in this backward process new numbers are produced only by reversing merges, since reversing an insertion simply means deleting one of the numbers. Therefore, all numbers appearing in the whole process will receive one of the two colors.

Sir Alex's first move is an insertion. Without loss of generality, assume this first number inserted is blue. Then, from this point on, until the last move, there is always at least one cell with a blue number.

Besides the last move, there is no move involving a blue and a red number, since all merges involves numbers with the same color, and insertions involve only one number. Call an insertion of a blue number or merge of two blue numbers a blue move, and define a red move analogously.

The whole sequence of blue moves could be repeated on another row of $k$ cells to produce one cell with the number $2^{n-1}$ and all the others empty, so there are at most $m(n-1, k)$ blue moves.

Now we look at the red moves. Since every time we perform a red move there is at least one cell occupied with a blue number, the whole sequence of red moves could be repeated on a row of $k-1$ cells to produce one cell with the number $2^{n-1}$ and all the others empty, so there are at most $m(n-1, k-1)$ red moves. This proves that

$$
m(n, k) \leqslant m(n-1, k)+m(n-1, k-1)+1 .
$$

On the other hand, we can start with an empty row of $k$ cells and perform $m(n-1, k)$ moves to produce one cell with the number $2^{n-1}$ and all the others empty, and after that perform $m(n-1, k-1)$ moves on those $k-1$ empty cells to produce the number $2^{n-1}$ in one of them, leaving $k-2$ empty. With one more merge we get one cell with $2^{n}$ and the others empty, proving that

$$
m(n, k) \geqslant m(n-1, k)+m(n-1, k-1)+1 .
$$

It follows that

$$
\begin{equation*}
m(n, k)=m(n-1, k)+m(n-1, k-1)+1, \tag{1}
\end{equation*}
$$

for $n \geqslant 1$ and $k \geqslant 2$.
If $k=1$ or $n=0$, we must insert $2^{n}$ on our first move and immediately get the final configuration, so $m(0, k)=1$ and $m(n, 1)=1$, for $n \geqslant 0$ and $k \geqslant 1$. These initial values, together with the recurrence relation (1), determine $m(n, k)$ uniquely.

Finally, we show that

$$
\begin{equation*}
m(n, k)=2 \sum_{j=0}^{k-1}\binom{n}{j}-1, \tag{2}
\end{equation*}
$$

for all integers $n \geqslant 0$ and $k \geqslant 1$.
We use induction on $n$. Since $m(0, k)=1$ for $k \geqslant 1,(2)$ is true for the base case. We make the induction hypothesis that (2) is true for some fixed positive integer $n$ and all $k \geqslant 1$. We have $m(n+1,1)=1=2\binom{n+1}{0}-1$, and for $k \geqslant 2$ the recurrence relation (1) and the induction hypothesis give us

$$
\begin{aligned}
& m(n+1, k)=m(n, k)+m(n, k-1)+1=2 \sum_{j=0}^{k-1}\binom{n}{j}-1+2 \sum_{j=0}^{k-2}\binom{n}{j}-1+1 \\
& \quad=2 \sum_{j=0}^{k-1}\binom{n}{j}+2 \sum_{j=0}^{k-1}\binom{n}{j-1}-1=2 \sum_{j=0}^{k-1}\left(\binom{n}{j}+\binom{n}{j-1}\right)-1=2 \sum_{j=0}^{k-1}\binom{n+1}{j}-1,
\end{aligned}
$$

which completes the proof.

Comment 1. After deducing the recurrence relation (1), it may be convenient to homogenize the recurrence relation by defining $h(n, k)=m(n, k)+1$. We get the new relation

$$
\begin{equation*}
h(n, k)=h(n-1, k)+h(n-1, k), \tag{3}
\end{equation*}
$$

for $n \geqslant 1$ and $k \geqslant 2$, with initial values $h(0, k)=h(n, 1)=2$, for $n \geqslant 0$ and $k \geqslant 1$.
This may help one to guess the answer, and also with other approaches like the one we develop next.

Comment 2. We can use a generating function to find the answer without guessing. We work with the homogenized recurrence relation (3). Define $h(n, 0)=0$ so that (3) is valid for $k=1$ as well. Now we set up the generating function $f(x, y)=\sum_{n, k \geqslant 0} h(n, k) x^{n} y^{k}$. Multiplying the recurrence relation (3) by $x^{n} y^{k}$ and summing over $n, k \geqslant 1$, we get

$$
\sum_{n, k \geqslant 1} h(n, k) x^{n} y^{k}=x \sum_{n, k \geqslant 1} h(n-1, k) x^{n-1} y^{k}+x y \sum_{n, k \geqslant 1} h(n-1, k-1) x^{n-1} y^{k-1} .
$$

Completing the missing terms leads to the following equation on $f(x, y)$ :

$$
f(x, y)-\sum_{n \geqslant 0} h(n, 0) x^{n}-\sum_{k \geqslant 1} h(0, k) y^{k}=x f(x, y)-x \sum_{n \geqslant 0} h(n, 0) x^{n}+x y f(x, y) .
$$

Substituting the initial values, we obtain

$$
f(x, y)=\frac{2 y}{1-y} \cdot \frac{1}{1-x(1+y)} .
$$

Developing as a power series, we get

$$
f(x, y)=2 \sum_{j \geqslant 1} y^{j} \cdot \sum_{n \geqslant 0}(1+y)^{n} x^{n} .
$$

The coefficient of $x^{n}$ in this power series is

$$
2 \sum_{j \geqslant 1} y^{j} \cdot(1+y)^{n}=2 \sum_{j \geqslant 1} y^{j} \cdot \sum_{i \geqslant 0}\binom{n}{i} y^{i},
$$

and extracting the coefficient of $y^{k}$ in this last expression we finally obtain the value for $h(n, k)$,

$$
h(n, k)=2 \sum_{j=0}^{k-1}\binom{n}{j} .
$$

This proves that

$$
m(n, k)=2 \sum_{j=0}^{k-1}\binom{n}{j}-1
$$

The generating function approach also works if applied to the non-homogeneous recurrence relation (1), but the computations are less straightforward.
Solution 2. Define merges and insertions as in Solution 1. After each move made by Sir Alex we compute the number $N$ of empty cells, and the sum $S$ of all the numbers written in the cells. Insertions always increase $S$ by some power of 2 , and increase $N$ exactly by 1 . Merges do not change $S$ and decrease $N$ exactly by 1 . Since the initial value of $N$ is 0 and its final value is 1 , the total number of insertions exceeds that of merges by exactly one. So, to maximize the number of moves, we need to maximize the number of insertions.

We will need the following lemma.
Lemma. If the binary representation of a positive integer $A$ has $d$ nonzero digits, then $A$ cannot be represented as a sum of fewer than $d$ powers of 2 . Moreover, any representation of $A$ as a sum of $d$ powers of 2 must coincide with its binary representation.
Proof. Let $s$ be the minimum number of summands in all possible representations of $A$ as sum of powers of 2 . Suppose there is such a representation with $s$ summands, where two of the summands are equal to each other. Then, replacing those two summands with the result of their sum, we obtain a representation with fewer than $s$ summands, which is a contradiction. We deduce that in any representation with $s$ summands, the summands are all distinct, so any such representation must coincide with the unique binary representation of $A$, and $s=d$.

Now we split the solution into a sequence of claims.
Claim 1. After every move, the number $S$ is the sum of at most $k-1$ distinct powers of 2 .
Proof. If $S$ is the sum of $k$ (or more) distinct powers of 2 , the Lemma implies that the $k$ cells are filled with these numbers. This is a contradiction since no more merges or insertions can be made.

Let $A(n, k-1)$ denote the set of all positive integers not exceeding $2^{n}$ with at most $k-1$ nonzero digits in its base 2 representation. Since every insertion increases the value of $S$, by Claim 1, the total number of insertions is at most $|A(n, k-1)|$. We proceed to prove that it is possible to achieve this number of insertions.
Claim 2. Let $A(n, k-1)=\left\{a_{1}, a_{2}, \ldots, a_{m}\right\}$, with $a_{1}<a_{2}<\cdots<a_{m}$. If after some of Sir Alex's moves the value of $S$ is $a_{j}$, with $j \in\{1,2, \ldots, m-1\}$, then there is a sequence of moves after which the value of $S$ is exactly $a_{j+1}$.
Proof. Suppose $S=a_{j}$. Performing all possible merges, we eventually get different powers of 2 in all nonempty cells. After that, by Claim 1 there will be at least one empty cell, in which we want to insert $a_{j+1}-a_{j}$. It remains to show that $a_{j+1}-a_{j}$ is a power of 2 .

For this purpose, we notice that if $a_{j}$ has less than $k-1$ nonzero digits in base 2 then $a_{j+1}=a_{j}+1$. Otherwise, we have $a_{j}=2^{b_{k-1}}+\cdots+2^{b_{2}}+2^{b_{1}}$ with $b_{1}<b_{2}<\cdots<b_{k-1}$. Then, adding any number less than $2^{b_{1}}$ to $a_{j}$ will result in a number with more than $k-1$ nonzero
binary digits. On the other hand, $a_{j}+2^{b_{1}}$ is a sum of $k$ powers of 2 , not all distinct, so by the Lemma it will be a sum of less then $k$ distinct powers of 2 . This means that $a_{j+1}-a_{j}=2^{b_{1}}$, completing the proof.

Claims 1 and 2 prove that the maximum number of insertions is $|A(n, k-1)|$. We now compute this number.
Claim 3. $|A(n, k-1)|=\sum_{j=0}^{k-1}\binom{n}{j}$.
Proof. The number $2^{n}$ is the only element of $A(n, k-1)$ with $n+1$ binary digits. Any other element has at most $n$ binary digits, at least one and at most $k-1$ of them are nonzero (so they are ones). For each $j \in\{1,2, \ldots, k-1\}$, there are $\binom{n}{j}$ such elements with exactly $j$ binary digits equal to one. We conclude that $|A(n, k-1)|=1+\sum_{j=1}^{k-1}\binom{n}{j}=\sum_{j=0}^{k-1}\binom{n}{j}$.

Recalling that the number of insertions exceeds that of merges by exactly 1 , we deduce that the maximum number of moves is $2 \sum_{j=0}^{k-1}\binom{n}{j}-1$.

C4. Let $N \geqslant 2$ be an integer. $N(N+1)$ soccer players, no two of the same height, stand in a row in some order. Coach Ralph wants to remove $N(N-1)$ people from this row so that in the remaining row of $2 N$ players, no one stands between the two tallest ones, no one stands between the third and the fourth tallest ones, ..., and finally no one stands between the two shortest ones. Show that this is always possible.

Solution 1. Split the row into $N$ blocks with $N+1$ consecutive people each. We will show how to remove $N-1$ people from each block in order to satisfy the coach's wish.

First, construct a $(N+1) \times N$ matrix where $x_{i, j}$ is the height of the $i^{\text {th }}$ tallest person of the $j^{\text {th }}$ block-in other words, each column lists the heights within a single block, sorted in decreasing order from top to bottom.

We will reorder this matrix by repeatedly swapping whole columns. First, by column permutation, make sure that $x_{2,1}=\max \left\{x_{2, i}: i=1,2, \ldots, N\right\}$ (the first column contains the largest height of the second row). With the first column fixed, permute the other ones so that $x_{3,2}=\max \left\{x_{3, i}: i=2, \ldots, N\right\}$ (the second column contains the tallest person of the third row, first column excluded). In short, at step $k(k=1,2, \ldots, N-1)$, we permute the columns from $k$ to $N$ so that $x_{k+1, k}=\max \left\{x_{i, k}: i=k, k+1, \ldots, N\right\}$, and end up with an array like this:

| $\boldsymbol{x}_{\mathbf{1 , 1}}$ | $x_{1,2}$ | $x_{1,3}$ | $\cdots$ | $x_{1, N-1}$ | $x_{1, N}$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| V | $\vee$ | $\vee$ |  | $\vee$ | $\vee$ |
| $\boldsymbol{x}_{\mathbf{2 , \mathbf { 1 }}}>$ | $\boldsymbol{x}_{\mathbf{2 , \mathbf { 2 }}}$ | $x_{2,3}$ | $\cdots$ | $x_{2, N-1}$ | $x_{2, N}$ |
| $\vee$ | $\vee$ | $\vee$ |  | $\vee$ | $\vee$ |
| $x_{3,1}$ | $\boldsymbol{x}_{\mathbf{3 , 2}}$ | $\boldsymbol{\boldsymbol { x } _ { \mathbf { 3 , 3 } }}$ | $\cdots$ | $x_{3, N-1}$ | $x_{3, N}$ |
| $\vee$ | $\vee$ | $\vee$ |  | $\vee$ | $\vee$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\ddots$ | $\vdots$ | $\vdots$ |
| $\vee$ | $\vee$ | $\vee$ |  | $\vee$ | $\vee$ |
| $x_{N, 1}$ | $x_{N, 2}$ | $x_{N, 3}$ | $\cdots$ | $\boldsymbol{x}_{\boldsymbol{N}, \boldsymbol{N}-\mathbf{1}}>$ | $\boldsymbol{x}_{\boldsymbol{N}, \boldsymbol{N}}$ |
| $\vee$ | $\vee$ | $\vee$ |  | $\vee$ | $\vee$ |
| $x_{N+1,1}$ | $x_{N+1,2}$ | $x_{N+1,3} \cdots$ | $x_{N+1, N-1}$ | $\boldsymbol{x}_{\boldsymbol{N + 1 , N}}$ |  |

Now we make the bold choice: from the original row of people, remove everyone but those with heights

$$
\begin{equation*}
x_{1,1}>x_{2,1}>x_{2,2}>x_{3,2}>\cdots>x_{N, N-1}>x_{N, N}>x_{N+1, N} \tag{*}
\end{equation*}
$$

Of course this height order (*) is not necessarily their spatial order in the new row. We now need to convince ourselves that each pair ( $x_{k, k} ; x_{k+1, k}$ ) remains spatially together in this new row. But $x_{k, k}$ and $x_{k+1, k}$ belong to the same column/block of consecutive $N+1$ people; the only people that could possibly stand between them were also in this block, and they are all gone.

Solution 2. Split the people into $N$ groups by height: group $G_{1}$ has the $N+1$ tallest ones, group $G_{2}$ has the next $N+1$ tallest, and so on, up to group $G_{N}$ with the $N+1$ shortest people.

Now scan the original row from left to right, stopping as soon as you have scanned two people (consecutively or not) from the same group, say, $G_{i}$. Since we have $N$ groups, this must happen before or at the $(N+1)^{\text {th }}$ person of the row. Choose this pair of people, removing all the other people from the same group $G_{i}$ and also all people that have been scanned so far. The only people that could separate this pair's heights were in group $G_{i}$ (and they are gone); the only people that could separate this pair's positions were already scanned (and they are gone too).

We are now left with $N-1$ groups (all except $G_{i}$ ). Since each of them lost at most one person, each one has at least $N$ unscanned people left in the row. Repeat the scanning process from left to right, choosing the next two people from the same group, removing this group and
everyone scanned up to that point. Once again we end up with two people who are next to each other in the remaining row and whose heights cannot be separated by anyone else who remains (since the rest of their group is gone). After picking these 2 pairs, we still have $N-2$ groups with at least $N-1$ people each.

If we repeat the scanning process a total of $N$ times, it is easy to check that we will end up with 2 people from each group, for a total of $2 N$ people remaining. The height order is guaranteed by the grouping, and the scanning construction from left to right guarantees that each pair from a group stand next to each other in the final row. We are done.

Solution 3. This is essentially the same as solution 1, but presented inductively. The essence of the argument is the following lemma.
Lemma. Assume that we have $N$ disjoint groups of at least $N+1$ people in each, all people have distinct heights. Then one can choose two people from each group so that among the chosen people, the two tallest ones are in one group, the third and the fourth tallest ones are in one group, ..., and the two shortest ones are in one group.
Proof. Induction on $N \geqslant 1$; for $N=1$, the statement is trivial.
Consider now $N$ groups $G_{1}, \ldots, G_{N}$ with at least $N+1$ people in each for $N \geqslant 2$. Enumerate the people by $1,2, \ldots, N(N+1)$ according to their height, say, from tallest to shortest. Find the least $s$ such that two people among $1,2, \ldots, s$ are in one group (without loss of generality, say this group is $G_{N}$ ). By the minimality of $s$, the two mentioned people in $G_{N}$ are $s$ and some $i<s$.

Now we choose people $i$ and $s$ in $G_{N}$, forget about this group, and remove the people $1,2, \ldots, s$ from $G_{1}, \ldots, G_{N-1}$. Due to minimality of $s$ again, each of the obtained groups $G_{1}^{\prime}, \ldots, G_{N-1}^{\prime}$ contains at least $N$ people. By the induction hypothesis, one can choose a pair of people from each of $G_{1}^{\prime}, \ldots, G_{N-1}^{\prime}$ so as to satisfy the required conditions. Since all these people have numbers greater than $s$, addition of the pair $(s, i)$ from $G_{N}$ does not violate these requirements.

To solve the problem, it suffices now to split the row into $N$ contiguous groups with $N+1$ people in each and apply the Lemma to those groups.

Comment 1. One can identify each person with a pair of indices $(p, h)(p, h \in\{1,2, \ldots, N(N+1)\})$ so that the $p^{\text {th }}$ person in the row (say, from left to right) is the $h^{\text {th }}$ tallest person in the group. Say that $(a, b)$ separates $\left(x_{1}, y_{1}\right)$ and $\left(x_{2}, y_{2}\right)$ whenever $a$ is strictly between $x_{1}$ and $y_{1}$, or $b$ is strictly between $x_{2}$ and $y_{2}$. So the coach wants to pick $2 N$ people $\left(p_{i}, h_{i}\right)(i=1,2, \ldots, 2 N)$ such that no chosen person separates ( $p_{1}, h_{1}$ ) from ( $p_{2}, h_{2}$ ), no chosen person separates ( $p_{3}, h_{3}$ ) and ( $p_{4}, h_{4}$ ), and so on. This formulation reveals a duality between positions and heights. In that sense, solutions 1 and 2 are dual of each other.

Comment 2. The number $N(N+1)$ is sharp for $N=2$ and $N=3$, due to arrangements $1,5,3,4,2$ and $1,10,6,4,3,9,5,8,7,2,11$.

C5. A hunter and an invisible rabbit play a game in the Euclidean plane. The hunter's starting point $H_{0}$ coincides with the rabbit's starting point $R_{0}$. In the $n^{\text {th }}$ round of the game ( $n \geqslant 1$ ), the following happens.
(1) First the invisible rabbit moves secretly and unobserved from its current point $R_{n-1}$ to some new point $R_{n}$ with $R_{n-1} R_{n}=1$.
(2) The hunter has a tracking device (e.g. dog) that returns an approximate position $R_{n}^{\prime}$ of the rabbit, so that $R_{n} R_{n}^{\prime} \leqslant 1$.
(3) The hunter then visibly moves from point $H_{n-1}$ to a new point $H_{n}$ with $H_{n-1} H_{n}=1$.

Is there a strategy for the hunter that guarantees that after $10^{9}$ such rounds the distance between the hunter and the rabbit is below 100 ?
(Austria)
Answer: There is no such strategy for the hunter. The rabbit "wins".
Solution. If the answer were "yes", the hunter would have a strategy that would "work", no matter how the rabbit moved or where the radar pings $R_{n}^{\prime}$ appeared. We will show the opposite: with bad luck from the radar pings, there is no strategy for the hunter that guarantees that the distance stays below 100 in $10^{9}$ rounds.

So, let $d_{n}$ be the distance between the hunter and the rabbit after $n$ rounds. Of course, if $d_{n} \geqslant 100$ for any $n<10^{9}$, the rabbit has won - it just needs to move straight away from the hunter, and the distance will be kept at or above 100 thereon.

We will now show that, while $d_{n}<100$, whatever given strategy the hunter follows, the rabbit has a way of increasing $d_{n}^{2}$ by at least $\frac{1}{2}$ every 200 rounds (as long as the radar pings are lucky enough for the rabbit). This way, $d_{n}^{2}$ will reach $10^{4}$ in less than $2 \cdot 10^{4} \cdot 200=4 \cdot 10^{6}<10^{9}$ rounds, and the rabbit wins.

Suppose the hunter is at $H_{n}$ and the rabbit is at $R_{n}$. Suppose even that the rabbit reveals its position at this moment to the hunter (this allows us to ignore all information from previous radar pings). Let $r$ be the line $H_{n} R_{n}$, and $Y_{1}$ and $Y_{2}$ be points which are 1 unit away from $r$ and 200 units away from $R_{n}$, as in the figure below.


The rabbit's plan is simply to choose one of the points $Y_{1}$ or $Y_{2}$ and hop 200 rounds straight towards it. Since all hops stay within 1 distance unit from $r$, it is possible that all radar pings stay on $r$. In particular, in this case, the hunter has no way of knowing whether the rabbit chose $Y_{1}$ or $Y_{2}$.

Looking at such pings, what is the hunter going to do? If the hunter's strategy tells him to go 200 rounds straight to the right, he ends up at point $H^{\prime}$ in the figure. Note that the hunter does not have a better alternative! Indeed, after these 200 rounds he will always end up at a point to the left of $H^{\prime}$. If his strategy took him to a point above $r$, he would end up even further from $Y_{2}$; and if his strategy took him below $r$, he would end up even further from $Y_{1}$. In other words, no matter what strategy the hunter follows, he can never be sure his distance to the rabbit will be less than $y \stackrel{\text { def }}{=} H^{\prime} Y_{1}=H^{\prime} Y_{2}$ after these 200 rounds.

To estimate $y^{2}$, we take $Z$ as the midpoint of segment $Y_{1} Y_{2}$, we take $R^{\prime}$ as a point 200 units to the right of $R_{n}$ and we define $\varepsilon=Z R^{\prime}$ (note that $H^{\prime} R^{\prime}=d_{n}$ ). Then

$$
y^{2}=1+\left(H^{\prime} Z\right)^{2}=1+\left(d_{n}-\varepsilon\right)^{2}
$$

where

$$
\varepsilon=200-R_{n} Z=200-\sqrt{200^{2}-1}=\frac{1}{200+\sqrt{200^{2}-1}}>\frac{1}{400} .
$$

In particular, $\varepsilon^{2}+1=400 \varepsilon$, so

$$
y^{2}=d_{n}^{2}-2 \varepsilon d_{n}+\varepsilon^{2}+1=d_{n}^{2}+\varepsilon\left(400-2 d_{n}\right) .
$$

Since $\varepsilon>\frac{1}{400}$ and we assumed $d_{n}<100$, this shows that $y^{2}>d_{n}^{2}+\frac{1}{2}$. So, as we claimed, with this list of radar pings, no matter what the hunter does, the rabbit might achieve $d_{n+200}^{2}>d_{n}^{2}+\frac{1}{2}$. The wabbit wins.

Comment 1. Many different versions of the solution above can be found by replacing 200 with some other number $N$ for the number of hops the rabbit takes between reveals. If this is done, we have:

$$
\varepsilon=N-\sqrt{N^{2}-1}>\frac{1}{N+\sqrt{N^{2}-1}}>\frac{1}{2 N}
$$

and

$$
\varepsilon^{2}+1=2 N \varepsilon,
$$

so, as long as $N>d_{n}$, we would find

$$
y^{2}=d_{n}^{2}+\varepsilon\left(2 N-2 d_{n}\right)>d_{n}^{2}+\frac{N-d_{n}}{N} .
$$

For example, taking $N=101$ is already enough-the squared distance increases by at least $\frac{1}{101}$ every 101 rounds, and $101^{2} \cdot 10^{4}=1.0201 \cdot 10^{8}<10^{9}$ rounds are enough for the rabbit. If the statement is made sharper, some such versions might not work any longer.

Comment 2. The original statement asked whether the distance could be kept under $10^{10}$ in $10^{100}$ rounds.

C6. Let $n>1$ be an integer. An $n \times n \times n$ cube is composed of $n^{3}$ unit cubes. Each unit cube is painted with one color. For each $n \times n \times 1$ box consisting of $n^{2}$ unit cubes (of any of the three possible orientations), we consider the set of the colors present in that box (each color is listed only once). This way, we get $3 n$ sets of colors, split into three groups according to the orientation. It happens that for every set in any group, the same set appears in both of the other groups. Determine, in terms of $n$, the maximal possible number of colors that are present.
(Russia)
Answer: The maximal number is $\frac{n(n+1)(2 n+1)}{6}$.
Solution 1. Call a $n \times n \times 1$ box an $x$-box, a $y$-box, or a $z$-box, according to the direction of its short side. Let $C$ be the number of colors in a valid configuration. We start with the upper bound for $C$.

Let $\mathcal{C}_{1}, \mathcal{C}_{2}$, and $\mathcal{C}_{3}$ be the sets of colors which appear in the big cube exactly once, exactly twice, and at least thrice, respectively. Let $M_{i}$ be the set of unit cubes whose colors are in $\mathcal{C}_{i}$, and denote $n_{i}=\left|M_{i}\right|$.

Consider any $x$-box $X$, and let $Y$ and $Z$ be a $y$ - and a $z$-box containing the same set of colors as $X$ does.
Claim. $4\left|X \cap M_{1}\right|+\left|X \cap M_{2}\right| \leqslant 3 n+1$.
Proof. We distinguish two cases.
Case 1: $X \cap M_{1} \neq \varnothing$.
A cube from $X \cap M_{1}$ should appear in all three boxes $X, Y$, and $Z$, so it should lie in $X \cap Y \cap Z$. Thus $X \cap M_{1}=X \cap Y \cap Z$ and $\left|X \cap M_{1}\right|=1$.

Consider now the cubes in $X \cap M_{2}$. There are at most 2( $n-1$ ) of them lying in $X \cap Y$ or $X \cap Z$ (because the cube from $X \cap Y \cap Z$ is in $M_{1}$ ). Let $a$ be some other cube from $X \cap M_{2}$. Recall that there is just one other cube $a^{\prime}$ sharing a color with $a$. But both $Y$ and $Z$ should contain such cube, so $a^{\prime} \in Y \cap Z$ (but $a^{\prime} \notin X \cap Y \cap Z$ ). The map $a \mapsto a^{\prime}$ is clearly injective, so the number of cubes $a$ we are interested in does not exceed $|(Y \cap Z) \backslash X|=n-1$. Thus $\left|X \cap M_{2}\right| \leqslant 2(n-1)+(n-1)=3(n-1)$, and hence $4\left|X \cap M_{1}\right|+\left|X \cap M_{2}\right| \leqslant 4+3(n-1)=3 n+1$. Case 2: $X \cap M_{1}=\varnothing$.

In this case, the same argument applies with several changes. Indeed, $X \cap M_{2}$ contains at most $2 n-1$ cubes from $X \cap Y$ or $X \cap Z$. Any other cube $a$ in $X \cap M_{2}$ corresponds to some $a^{\prime} \in Y \cap Z$ (possibly with $a^{\prime} \in X$ ), so there are at most $n$ of them. All this results in $\left|X \cap M_{2}\right| \leqslant(2 n-1)+n=3 n-1$, which is even better than we need (by the assumptions of our case).

Summing up the inequalities from the Claim over all $x$-boxes $X$, we obtain

$$
4 n_{1}+n_{2} \leqslant n(3 n+1)
$$

Obviously, we also have $n_{1}+n_{2}+n_{3}=n^{3}$.
Now we are prepared to estimate $C$. Due to the definition of the $M_{i}$, we have $n_{i} \geqslant i\left|\mathcal{C}_{i}\right|$, so

$$
C \leqslant n_{1}+\frac{n_{2}}{2}+\frac{n_{3}}{3}=\frac{n_{1}+n_{2}+n_{3}}{3}+\frac{4 n_{1}+n_{2}}{6} \leqslant \frac{n^{3}}{3}+\frac{3 n^{2}+n}{6}=\frac{n(n+1)(2 n+1)}{6} .
$$

It remains to present an example of an appropriate coloring in the above-mentioned number of colors. For each color, we present the set of all cubes of this color. These sets are:

1. $n$ singletons of the form $S_{i}=\{(i, i, i)\}$ (with $1 \leqslant i \leqslant n$ );
2. $3\binom{n}{2}$ doubletons of the forms $D_{i, j}^{1}=\{(i, j, j),(j, i, i)\}, D_{i, j}^{2}=\{(j, i, j),(i, j, i)\}$, and $D_{i, j}^{3}=$ $\{(j, j, i),(i, i, j)\}$ (with $1 \leqslant i<j \leqslant n)$;
3. $2\binom{n}{3}$ triplets of the form $T_{i, j, k}=\{(i, j, k),(j, k, i),(k, i, j)\}$ (with $1 \leqslant i<j<k \leqslant n$ or $1 \leqslant i<k<j \leqslant n)$.

One may easily see that the $i^{\text {th }}$ boxes of each orientation contain the same set of colors, and that

$$
n+\frac{3 n(n-1)}{2}+\frac{n(n-1)(n-2)}{3}=\frac{n(n+1)(2 n+1)}{6}
$$

colors are used, as required.
Solution 2. We will approach a new version of the original problem. In this new version, each cube may have a color, or be invisible (not both). Now we make sets of colors for each $n \times n \times 1$ box as before (where "invisible" is not considered a color) and group them by orientation, also as before. Finally, we require that, for every non-empty set in any group, the same set must appear in the other 2 groups. What is the maximum number of colors present with these new requirements?

Let us call strange a big $n \times n \times n$ cube whose painting scheme satisfies the new requirements, and let $D$ be the number of colors in a strange cube. Note that any cube that satisfies the original requirements is also strange, so $\max (D)$ is an upper bound for the original answer.
Claim. $D \leqslant \frac{n(n+1)(2 n+1)}{6}$.
Proof. The proof is by induction on $n$. If $n=1$, we must paint the cube with at most 1 color.
Now, pick a $n \times n \times n$ strange cube $A$, where $n \geqslant 2$. If $A$ is completely invisible, $D=0$ and we are done. Otherwise, pick a non-empty set of colors $\mathcal{S}$ which corresponds to, say, the boxes $X, Y$ and $Z$ of different orientations.

Now find all cubes in $A$ whose colors are in $\mathcal{S}$ and make them invisible. Since $X, Y$ and $Z$ are now completely invisible, we can throw them away and focus on the remaining $(n-1) \times(n-1) \times(n-1)$ cube $B$. The sets of colors in all the groups for $B$ are the same as the sets for $A$, removing exactly the colors in $\mathcal{S}$, and no others! Therefore, every nonempty set that appears in one group for $B$ still shows up in all possible orientations (it is possible that an empty set of colors in $B$ only matched $X, Y$ or $Z$ before these were thrown away, but remember we do not require empty sets to match anyway). In summary, $B$ is also strange.

By the induction hypothesis, we may assume that $B$ has at most $\frac{(n-1) n(2 n-1)}{6}$ colors. Since there were at most $n^{2}$ different colors in $\mathcal{S}$, we have that $A$ has at most $\frac{(n-1) n(2 n-1)}{6}+n^{2}=$ $\frac{n(n+1)(2 n+1)}{6}$ colors.

Finally, the construction in the previous solution shows a painting scheme (with no invisible cubes) that reaches this maximum, so we are done.

C7. For any finite sets $X$ and $Y$ of positive integers, denote by $f_{X}(k)$ the $k^{\text {th }}$ smallest positive integer not in $X$, and let

$$
X * Y=X \cup\left\{f_{X}(y): y \in Y\right\}
$$

Let $A$ be a set of $a>0$ positive integers, and let $B$ be a set of $b>0$ positive integers. Prove that if $A * B=B * A$, then

$$
\underbrace{A *(A * \cdots *(A *(A * A)) \cdots)}_{A \text { appears } b \text { times }}=\underbrace{B *(B * \cdots *(B *(B * B)) \cdots)}_{B \text { appears } a \text { times }}
$$

(U.S.A.)

Solution 1. For any function $g: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ and any subset $X \subset \mathbb{Z}_{>0}$, we define $g(X)=$ $\{g(x): x \in X\}$. We have that the image of $f_{X}$ is $f_{X}\left(\mathbb{Z}_{>0}\right)=\mathbb{Z}_{>0} \backslash X$. We now show a general lemma about the operation *, with the goal of showing that $*$ is associative.
Lemma 1. Let $X$ and $Y$ be finite sets of positive integers. The functions $f_{X * Y}$ and $f_{X} \circ f_{Y}$ are equal.
Proof. We have
$f_{X * Y}\left(\mathbb{Z}_{>0}\right)=\mathbb{Z}_{>0} \backslash(X * Y)=\left(\mathbb{Z}_{>0} \backslash X\right) \backslash f_{X}(Y)=f_{X}\left(\mathbb{Z}_{>0}\right) \backslash f_{X}(Y)=f_{X}\left(\mathbb{Z}_{>0} \backslash Y\right)=f_{X}\left(f_{Y}\left(\mathbb{Z}_{>0}\right)\right)$.
Thus, the functions $f_{X * Y}$ and $f_{X} \circ f_{Y}$ are strictly increasing functions with the same range. Because a strictly function is uniquely defined by its range, we have $f_{X * Y}=f_{X} \circ f_{Y}$.

Lemma 1 implies that * is associative, in the sense that $(A * B) * C=A *(B * C)$ for any finite sets $A, B$, and $C$ of positive integers. We prove the associativity by noting

$$
\begin{gathered}
\mathbb{Z}_{>0} \backslash((A * B) * C)=f_{(A * B) * C}\left(\mathbb{Z}_{>0}\right)=f_{A * B}\left(f_{C}\left(\mathbb{Z}_{>0}\right)\right)=f_{A}\left(f_{B}\left(f_{C}\left(\mathbb{Z}_{>0}\right)\right)\right) \\
=f_{A}\left(f_{B * C}\left(\mathbb{Z}_{>0}\right)=f_{A *(B * C)}\left(\mathbb{Z}_{>0}\right)=\mathbb{Z}_{>0} \backslash(A *(B * C)) .\right.
\end{gathered}
$$

In light of the associativity of *, we may drop the parentheses when we write expressions like $A *(B * C)$. We also introduce the notation

$$
X^{* k}=\underbrace{X *(X * \cdots *(X *(X * X)) \ldots)}_{X \text { appears } k \text { times }}
$$

Our goal is then to show that $A * B=B * A$ implies $A^{* b}=B^{* a}$. We will do so via the following general lemma.
Lemma 2. Suppose that $X$ and $Y$ are finite sets of positive integers satisfying $X * Y=Y * X$ and $|X|=|Y|$. Then, we must have $X=Y$.
Proof. Assume that $X$ and $Y$ are not equal. Let $s$ be the largest number in exactly one of $X$ and $Y$. Without loss of generality, say that $s \in X \backslash Y$. The number $f_{X}(s)$ counts the $s^{t h}$ number not in $X$, which implies that

$$
\begin{equation*}
f_{X}(s)=s+\left|X \cap\left\{1,2, \ldots, f_{X}(s)\right\}\right| \tag{1}
\end{equation*}
$$

Since $f_{X}(s) \geqslant s$, we have that

$$
\left\{f_{X}(s)+1, f_{X}(s)+2, \ldots\right\} \cap X=\left\{f_{X}(s)+1, f_{X}(s)+2, \ldots\right\} \cap Y
$$

which, together with the assumption that $|X|=|Y|$, gives

$$
\begin{equation*}
\left|X \cap\left\{1,2, \ldots, f_{X}(s)\right\}\right|=\left|Y \cap\left\{1,2, \ldots, f_{X}(s)\right\}\right| \tag{2}
\end{equation*}
$$

Now consider the equation

$$
t-|Y \cap\{1,2, \ldots, t\}|=s
$$

This equation is satisfied only when $t \in\left[f_{Y}(s), f_{Y}(s+1)\right)$, because the left hand side counts the number of elements up to $t$ that are not in $Y$. We have that the value $t=f_{X}(s)$ satisfies the above equation because of (1) and (2). Furthermore, since $f_{X}(s) \notin X$ and $f_{X}(s) \geqslant s$, we have that $f_{X}(s) \notin Y$ due to the maximality of $s$. Thus, by the above discussion, we must have $f_{X}(s)=f_{Y}(s)$.

Finally, we arrive at a contradiction. The value $f_{X}(s)$ is neither in $X$ nor in $f_{X}(Y)$, because $s$ is not in $Y$ by assumption. Thus, $f_{X}(s) \notin X * Y$. However, since $s \in X$, we have $f_{Y}(s) \in Y * X$, a contradiction.

We are now ready to finish the proof. Note first of all that $\left|A^{* b}\right|=a b=\left|B^{* a}\right|$. Moreover, since $A * B=B * A$, and * is associative, it follows that $A^{* b} * B^{* a}=B^{* a} * A^{* b}$. Thus, by Lemma 2, we have $A^{* b}=B^{* a}$, as desired.

Comment 1. Taking $A=X^{* k}$ and $B=X^{* l}$ generates many non-trivial examples where $A * B=B * A$. There are also other examples not of this form. For example, if $A=\{1,2,4\}$ and $B=\{1,3\}$, then $A * B=\{1,2,3,4,6\}=B * A$.
Solution 2. We will use Lemma 1 from Solution 1. Additionally, let $X^{* k}$ be defined as in Solution 1. If $X$ and $Y$ are finite sets, then

$$
\begin{equation*}
f_{X}=f_{Y} \Longleftrightarrow f_{X}\left(\mathbb{Z}_{>0}\right)=f_{Y}\left(\mathbb{Z}_{>0}\right) \Longleftrightarrow\left(\mathbb{Z}_{>0} \backslash X\right)=\left(\mathbb{Z}_{>0} \backslash Y\right) \Longleftrightarrow X=Y, \tag{3}
\end{equation*}
$$

where the first equivalence is because $f_{X}$ and $f_{Y}$ are strictly increasing functions, and the second equivalence is because $f_{X}\left(\mathbb{Z}_{>0}\right)=\mathbb{Z}_{>0} \backslash X$ and $f_{Y}\left(\mathbb{Z}_{>0}\right)=\mathbb{Z}_{>0} \backslash Y$.

Denote $g=f_{A}$ and $h=f_{B}$. The given relation $A * B=B * A$ is equivalent to $f_{A * B}=f_{B * A}$ because of (3), and by Lemma 1 of the first solution, this is equivalent to $g \circ h=h \circ g$. Similarly, the required relation $A^{* b}=B^{* a}$ is equivalent to $g^{b}=h^{a}$. We will show that

$$
\begin{equation*}
g^{b}(n)=h^{a}(n) \tag{4}
\end{equation*}
$$

for all $n \in \mathbb{Z}_{>0}$, which suffices to solve the problem.
To start, we claim that (4) holds for all sufficiently large $n$. Indeed, let $p$ and $q$ be the maximal elements of $A$ and $B$, respectively; we may assume that $p \geqslant q$. Then, for every $n \geqslant p$ we have $g(n)=n+a$ and $h(n)=n+b$, whence $g^{b}(n)=n+a b=h^{a}(n)$, as was claimed.

In view of this claim, if (4) is not identically true, then there exists a maximal $s$ with $g^{b}(s) \neq$ $h^{a}(s)$. Without loss of generality, we may assume that $g(s) \neq s$, for if we had $g(s)=h(s)=s$, then $s$ would satisfy (4). As $g$ is increasing, we then have $g(s)>s$, so (4) holds for $n=g(s)$. But then we have

$$
g\left(g^{b}(s)\right)=g^{b+1}(s)=g^{b}(n)=h^{a}(n)=h^{a}(g(s))=g\left(h^{a}(s)\right),
$$

where the last equality holds in view of $g \circ h=h \circ g$. By the injectivity of $g$, the above equality yields $g^{b}(s)=h^{a}(s)$, which contradicts the choice of $s$. Thus, we have proved that (4) is identically true on $\mathbb{Z}_{>0}$, as desired.

Comment 2. We present another proof of Lemma 2 of the first solution.
Let $x=|X|=|Y|$. Say that $u$ is the smallest number in $X$ and $v$ is the smallest number in $Y$; assume without loss of generality that $u \leqslant v$.

Let $T$ be any finite set of positive integers, and define $t=|T|$. Enumerate the elements of $X$ as $x_{1}<x_{2}<\cdots<x_{n}$. Define $S_{m}=f_{\left(T * X^{*(m-1)}\right)}(X)$, and enumerate its elements $s_{m, 1}<s_{m, 2}<\cdots<$ $s_{m, n}$. Note that the $S_{m}$ are pairwise disjoint; indeed, if we have $m<m^{\prime}$, then

$$
S_{m} \subset T * X^{* m} \subset T * X^{*\left(m^{\prime}-1\right)} \quad \text { and } \quad S_{m^{\prime}}=\left(T * X^{* m^{\prime}}\right) \backslash\left(T * X^{*\left(m^{\prime}-1\right)}\right)
$$

We claim the following statement, which essentially says that the $S_{m}$ are eventually linear translates of each other:

Claim. For every $i$, there exists some $m_{i}$ and $c_{i}$ such that for all $m>m_{i}$, we have that $s_{m, i}=t+m n-c_{i}$. Furthermore, the $c_{i}$ do not depend on the choice of $T$.

First, we show that this claim implies Lemma 2. We may choose $T=X$ and $T=Y$. Then, there is some $m^{\prime}$ such that for all $m \geqslant m^{\prime}$, we have

$$
\begin{equation*}
f_{X^{* m}}(X)=f_{\left(Y * X^{*(m-1)}\right)}(X) \tag{5}
\end{equation*}
$$

Because $u$ is the minimum element of $X, v$ is the minimum element of $Y$, and $u \leqslant v$, we have that

$$
\left(\bigcup_{m=m^{\prime}}^{\infty} f_{X * m}(X)\right) \cup X^{* m^{\prime}}=\left(\bigcup_{m=m^{\prime}}^{\infty} f_{\left(Y * X^{*(m-1)}\right)}(X)\right) \cup\left(Y * X^{*\left(m^{\prime}-1\right)}\right)=\{u, u+1, \ldots\},
$$

and in both the first and second expressions, the unions are of pairwise distinct sets. By (5), we obtain $X^{* m^{\prime}}=Y * X^{*\left(m^{\prime}-1\right)}$. Now, because $X$ and $Y$ commute, we get $X^{* m^{\prime}}=X^{*\left(m^{\prime}-1\right)} * Y$, and so $X=Y$.

We now prove the claim.
Proof of the claim. We induct downwards on $i$, first proving the statement for $i=n$, and so on.
Assume that $m$ is chosen so that all elements of $S_{m}$ are greater than all elements of $T$ (which is possible because $T$ is finite). For $i=n$, we have that $s_{m, n}>s_{k, n}$ for every $k<m$. Thus, all ( $m-1$ ) $n$ numbers of the form $s_{k, u}$ for $k<m$ and $1 \leqslant u \leqslant n$ are less than $s_{m, n}$. We then have that $s_{m, n}$ is the $\left((m-1) n+x_{n}\right)^{t h}$ number not in $T$, which is equal to $t+(m-1) n+x_{n}$. So we may choose $c_{n}=x_{n}-n$, which does not depend on $T$, which proves the base case for the induction.

For $i<n$, we have again that all elements $s_{m, j}$ for $j<i$ and $s_{p, i}$ for $p<m$ are less than $s_{m, i}$, so $s_{m, i}$ is the $\left((m-1) i+x_{i}\right)^{t h}$ element not in $T$ or of the form $s_{p, j}$ for $j>i$ and $p<m$. But by the inductive hypothesis, each of the sequences $s_{p, j}$ is eventually periodic with period $n$, and thus the sequence $s_{m, i}$ such must be as well. Since each of the sequences $s_{p, j}-t$ with $j>i$ eventually do not depend on $T$, the sequence $s_{m, i}-t$ eventually does not depend on $T$ either, so the inductive step is complete. This proves the claim and thus Lemma 2.

C8. Let $n$ be a given positive integer. In the Cartesian plane, each lattice point with nonnegative coordinates initially contains a butterfly, and there are no other butterflies. The neighborhood of a lattice point $c$ consists of all lattice points within the axis-aligned $(2 n+1) \times$ $(2 n+1)$ square centered at $c$, apart from $c$ itself. We call a butterfly lonely, crowded, or comfortable, depending on whether the number of butterflies in its neighborhood $N$ is respectively less than, greater than, or equal to half of the number of lattice points in $N$.

Every minute, all lonely butterflies fly away simultaneously. This process goes on for as long as there are any lonely butterflies. Assuming that the process eventually stops, determine the number of comfortable butterflies at the final state.
(Bulgaria)
Answer: $n^{2}+1$.
Solution. We always identify a butterfly with the lattice point it is situated at. For two points $p$ and $q$, we write $p \geqslant q$ if each coordinate of $p$ is at least the corresponding coordinate of $q$. Let $O$ be the origin, and let $\mathcal{Q}$ be the set of initially occupied points, i.e., of all lattice points with nonnegative coordinates. Let $\mathcal{R}_{\mathrm{H}}=\{(x, 0): x \geqslant 0\}$ and $\mathcal{R}_{\mathrm{V}}=\{(0, y): y \geqslant 0\}$ be the sets of the lattice points lying on the horizontal and vertical boundary rays of $\mathcal{Q}$. Denote by $N(a)$ the neighborhood of a lattice point $a$.

1. Initial observations. We call a set of lattice points up-right closed if its points stay in the set after being shifted by any lattice vector $(i, j)$ with $i, j \geqslant 0$. Whenever the butterflies form a up-right closed set $\mathcal{S}$, we have $|N(p) \cap \mathcal{S}| \geqslant|N(q) \cap \mathcal{S}|$ for any two points $p, q \in \mathcal{S}$ with $p \geqslant q$. So, since $\mathcal{Q}$ is up-right closed, the set of butterflies at any moment also preserves this property. We assume all forthcoming sets of lattice points to be up-right closed.

When speaking of some set $\mathcal{S}$ of lattice points, we call its points lonely, comfortable, or crowded with respect to this set (i.e., as if the butterflies were exactly at all points of $\mathcal{S}$ ). We call a set $\mathcal{S} \subset \mathcal{Q}$ stable if it contains no lonely points. In what follows, we are interested only in those stable sets whose complements in $\mathcal{Q}$ are finite, because one can easily see that only a finite number of butterflies can fly away on each minute.

If the initial set $\mathcal{Q}$ of butterflies contains some stable set $\mathcal{S}$, then, clearly no butterfly of this set will fly away. On the other hand, the set $\mathcal{F}$ of all butterflies in the end of the process is stable. This means that $\mathcal{F}$ is the largest (with respect to inclusion) stable set within $\mathcal{Q}$, and we are about to describe this set.
2. A description of a final set. The following notion will be useful. Let $\mathcal{U}=\left\{\vec{u}_{1}, \vec{u}_{2}, \ldots, \vec{u}_{d}\right\}$ be a set of $d$ pairwise non-parallel lattice vectors, each having a positive $x$ - and a negative $y$-coordinate. Assume that they are numbered in increasing order according to slope. We now define a $\mathcal{U}$-curve to be the broken line $p_{0} p_{1} \ldots p_{d}$ such that $p_{0} \in \mathcal{R}_{V}, p_{d} \in \mathcal{R}_{\mathrm{H}}$, and $\vec{p} i-1^{p_{i}}=\vec{u}_{i}$ for all $i=1,2, \ldots, m$ (see the Figure below to the left).


Construction of $\mathcal{U}$-curve


Construction of $\mathcal{D}$

Now, let $\mathcal{K}_{n}=\{(i, j): 1 \leqslant i \leqslant n,-n \leqslant j \leqslant-1\}$. Consider all the rays emerging at $O$ and passing through a point from $\mathcal{K}_{n}$; number them as $r_{1}, \ldots, r_{m}$ in increasing order according to slope. Let $A_{i}$ be the farthest from $O$ lattice point in $r_{i} \cap \mathcal{K}_{n}$, set $k_{i}=\left|r_{i} \cap \mathcal{K}_{n}\right|$, let $\vec{v}_{i}=\overrightarrow{O A_{i}}$, and finally denote $\mathcal{V}=\left\{\vec{v}_{i}: 1 \leqslant i \leqslant m\right\}$; see the Figure above to the right. We will concentrate on the $\mathcal{V}$-curve $d_{0} d_{1} \ldots d_{m}$; let $\mathcal{D}$ be the set of all lattice points $p$ such that $p \geqslant p^{\prime}$ for some (not necessarily lattice) point $p^{\prime}$ on the $\mathcal{V}$-curve. In fact, we will show that $\mathcal{D}=\mathcal{F}$.

Clearly, the $\mathcal{V}$-curve is symmetric in the line $y=x$. Denote by $D$ the convex hull of $\mathcal{D}$.
3. We prove that the set $\mathcal{D}$ contains all stable sets. Let $\mathcal{S} \subset \mathcal{Q}$ be a stable set (recall that it is assumed to be up-right closed and to have a finite complement in $\mathcal{Q}$ ). Denote by $S$ its convex hull; clearly, the vertices of $S$ are lattice points. The boundary of $S$ consists of two rays (horizontal and vertical ones) along with some $\mathcal{V}_{*}$-curve for some set of lattice vectors $\mathcal{V}_{*}$.
Claim 1. For every $\vec{v}_{i} \in \mathcal{V}$, there is a $\vec{v}_{i}^{*} \in \mathcal{V}_{*}$ co-directed with $\vec{v}$ with $\left|\vec{v}_{i}^{*}\right| \geqslant|\vec{v}|$.
Proof. Let $\ell$ be the supporting line of $S$ parallel to $\vec{v}_{i}$ (i.e., $\ell$ contains some point of $S$, and the set $S$ lies on one side of $\ell$ ). Take any point $b \in \ell \cap \mathcal{S}$ and consider $N(b)$. The line $\ell$ splits the set $N(b) \backslash \ell$ into two congruent parts, one having an empty intersection with $\mathcal{S}$. Hence, in order for $b$ not to be lonely, at least half of the set $\ell \cap N(b)$ (which contains $2 k_{i}$ points) should lie in $S$. Thus, the boundary of $S$ contains a segment $\ell \cap S$ with at least $k_{i}+1$ lattice points (including b) on it; this segment corresponds to the required vector $\vec{v}_{i}^{*} \in \mathcal{V}_{*}$.


Claim 2. Each stable set $\mathcal{S} \subseteq \mathcal{Q}$ lies in $\mathcal{D}$.
Proof. To show this, it suffices to prove that the $\mathcal{V}_{*}$-curve lies in $D$, i.e., that all its vertices do so. Let $p^{\prime}$ be an arbitrary vertex of the $\mathcal{V}_{*}$-curve; $p^{\prime}$ partitions this curve into two parts, $\mathcal{X}$ (being down-right of $p$ ) and $\mathcal{Y}$ (being up-left of $p$ ). The set $\mathcal{V}$ is split now into two parts: $\mathcal{V}_{\mathcal{X}}$ consisting of those $\vec{v}_{i} \in \mathcal{V}$ for which $\vec{v}_{i}^{*}$ corresponds to a segment in $\mathcal{X}$, and a similar part $\mathcal{V}_{\mathcal{Y}}$. Notice that the $\mathcal{V}$-curve consists of several segments corresponding to $\mathcal{V}_{\mathcal{X}}$, followed by those corresponding to $\mathcal{V}_{\mathcal{Y}}$. Hence there is a vertex $p$ of the $\mathcal{V}$-curve separating $\mathcal{V}_{\mathcal{X}}$ from $\mathcal{V}_{\mathcal{y}}$. Claim 1 now yields that $p^{\prime} \geqslant p$, so $p^{\prime} \in \mathcal{D}$, as required.

Claim 2 implies that the final set $\mathcal{F}$ is contained in $\mathcal{D}$.
4. $\mathcal{D}$ is stable, and its comfortable points are known. Recall the definitions of $r_{i}$; let $r_{i}^{\prime}$ be the ray complementary to $r_{i}$. By our definitions, the set $N(O)$ contains no points between the rays $r_{i}$ and $r_{i+1}$, as well as between $r_{i}^{\prime}$ and $r_{i+1}^{\prime}$.
Claim 3. In the set $\mathcal{D}$, all lattice points of the $\mathcal{V}$-curve are comfortable.
Proof. Let $p$ be any lattice point of the $\mathcal{V}$-curve, belonging to some segment $d_{i} d_{i+1}$. Draw the line $\ell$ containing this segment. Then $\ell \cap \mathcal{D}$ contains exactly $k_{i}+1$ lattice points, all of which lie in $N(p)$ except for $p$. Thus, exactly half of the points in $N(p) \cap \ell$ lie in $\mathcal{D}$. It remains to show that all points of $N(p)$ above $\ell$ lie in $\mathcal{D}$ (recall that all the points below $\ell$ lack this property).

Notice that each vector in $\mathcal{V}$ has one coordinate greater than $n / 2$; thus the neighborhood of $p$ contains parts of at most two segments of the $\mathcal{V}$-curve succeeding $d_{i} d_{i+1}$, as well as at most two of those preceding it.

The angles formed by these consecutive segments are obtained from those formed by $r_{j}$ and $r_{j-1}^{\prime}$ (with $i-1 \leqslant j \leqslant i+2$ ) by shifts; see the Figure below. All the points in $N(p)$ above $\ell$ which could lie outside $\mathcal{D}$ lie in shifted angles between $r_{j}, r_{j+1}$ or $r_{j}^{\prime}, r_{j-1}^{\prime}$. But those angles, restricted to $N(p)$, have no lattice points due to the above remark. The claim is proved.


Claim 4. All the points of $\mathcal{D}$ which are not on the boundary of $D$ are crowded.
Proof. Let $p \in \mathcal{D}$ be such a point. If it is to the up-right of some point $p^{\prime}$ on the curve, then the claim is easy: the shift of $N\left(p^{\prime}\right) \cap \mathcal{D}$ by $\overrightarrow{p^{\prime} p}$ is still in $\mathcal{D}$, and $N(p)$ contains at least one more point of $\mathcal{D}$ - either below or to the left of $p$. So, we may assume that $p$ lies in a right triangle constructed on some hypothenuse $d_{i} d_{i+1}$. Notice here that $d_{i}, d_{i+1} \in N(p)$.

Draw a line $\ell \| d_{i} d_{i+1}$ through $p$, and draw a vertical line $h$ through $d_{i}$; see Figure below. Let $\mathcal{D}_{\mathrm{L}}$ and $\mathcal{D}_{\mathrm{R}}$ be the parts of $\mathcal{D}$ lying to the left and to the right of $h$, respectively (points of $\mathcal{D} \cap h$ lie in both parts).


Notice that the vectors $\overrightarrow{d_{i} p}, \overrightarrow{d_{i+1} d_{i+2}}, \overrightarrow{d_{i} d_{i+1}}, \overrightarrow{d_{i-1} d_{i}}$, and $\overrightarrow{p d_{i+1}}$ are arranged in non-increasing order by slope. This means that $\mathcal{D}_{\mathrm{L}}$ shifted by $\overrightarrow{d_{i} p}$ still lies in $\mathcal{D}$, as well as $\mathcal{D}_{\mathrm{R}}$ shifted by $\overrightarrow{d_{i+1} p}$. As we have seen in the proof of Claim 3, these two shifts cover all points of $N(p)$ above $\ell$, along with those on $\ell$ to the left of $p$. Since $N(p)$ contains also $d_{i}$ and $d_{i+1}$, the point $p$ is crowded.

Thus, we have proved that $\mathcal{D}=\mathcal{F}$, and have shown that the lattice points on the $\mathcal{V}$-curve are exactly the comfortable points of $\mathcal{D}$. It remains to find their number.

Recall the definition of $\mathcal{K}_{n}$ (see Figure on the first page of the solution). Each segment $d_{i} d_{i+1}$ contains $k_{i}$ lattice points different from $d_{i}$. Taken over all $i$, these points exhaust all the lattice points in the $\mathcal{V}$-curve, except for $d_{1}$, and thus the number of lattice points on the $\mathcal{V}$-curve is $1+\sum_{i=1}^{m} k_{i}$. On the other hand, $\sum_{i=1}^{m} k_{i}$ is just the number of points in $\mathcal{K}_{n}$, so it equals $n^{2}$. Hence the answer to the problem is $n^{2}+1$.

Comment 1. The assumption that the process eventually stops is unnecessary for the problem, as one can see that, in fact, the process stops for every $n \geqslant 1$. Indeed, the proof of Claims 3 and 4 do not rely essentially on this assumption, and they together yield that the set $\mathcal{D}$ is stable. So, only butterflies that are not in $\mathcal{D}$ may fly away, and this takes only a finite time.

This assumption has been inserted into the problem statement in order to avoid several technical details regarding finiteness issues. It may also simplify several other arguments.

Comment 2. The description of the final set $\mathcal{F}(=\mathcal{D})$ seems to be crucial for the solution; the Problem Selection Committee is not aware of any solution that completely avoids such a description.

On the other hand, after the set $\mathcal{D}$ has been defined, the further steps may be performed in several ways. For example, in order to prove that all butterflies outside $\mathcal{D}$ will fly away, one may argue as follows. (Here we will also make use of the assumption that the process eventually stops.)

First of all, notice that the process can be modified in the following manner: Each minute, exactly one of the lonely butterflies flies away, until there are no more lonely butterflies. The modified process necessarily stops at the same state as the initial one. Indeed, one may observe, as in solution above, that the (unique) largest stable set is still the final set for the modified process.

Thus, in order to prove our claim, it suffices to indicate an order in which the butterflies should fly away in the new process; if we are able to exhaust the whole set $\mathcal{Q} \backslash \mathcal{D}$, we are done.

Let $\mathcal{C}_{0}=d_{0} d_{1} \ldots d_{m}$ be the $\mathcal{V}$-curve. Take its copy $\mathcal{C}$ and shift it downwards so that $d_{0}$ comes to some point below the origin $O$. Now we start moving $\mathcal{C}$ upwards continuously, until it comes back to its initial position $\mathcal{C}_{0}$. At each moment when $\mathcal{C}$ meets some lattice points, we convince all the butterflies at those points to fly away in a certain order. We will now show that we always have enough arguments for butterflies to do so, which will finish our argument for the claim..

Let $\mathcal{C}^{\prime}=d_{0}^{\prime} d_{1}^{\prime} \ldots d_{m}^{\prime}$ be a position of $\mathcal{C}$ when it meets some butterflies. We assume that all butterflies under this current position of $\mathcal{C}$ were already convinced enough and flied away. Consider the lowest butterfly $b$ on $\mathcal{C}^{\prime}$. Let $d_{i}^{\prime} d_{i+1}^{\prime}$ be the segment it lies on; we choose $i$ so that $b \neq d_{i+1}^{\prime}$ (this is possible because $\mathcal{C}$ as not yet reached $\mathcal{C}_{0}$ ).

Draw a line $\ell$ containing the segment $d_{i}^{\prime} d_{i+1}^{\prime}$. Then all the butterflies in $N(b)$ are situated on or above $\ell$; moreover, those on $\ell$ all lie on the segment $d_{i} d_{i+1}$. But this segment now contains at most $k_{i}$ butterflies (including b), since otherwise some butterfly had to occupy $d_{i+1}^{\prime}$ which is impossible by the choice of $b$. Thus, $b$ is lonely and hence may be convinced to fly away.

After $b$ has flied away, we switch to the lowest of the remaining butterflies on $\mathcal{C}^{\prime}$, and so on.
Claims 3 and 4 also allow some different proofs which are not presented here.

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## Geometry

G1. Let $A B C D E$ be a convex pentagon such that $A B=B C=C D, \angle E A B=\angle B C D$, and $\angle E D C=\angle C B A$. Prove that the perpendicular line from $E$ to $B C$ and the line segments $A C$ and $B D$ are concurrent.
(Italy)
Solution 1. Throughout the solution, we refer to $\angle A, \angle B, \angle C, \angle D$, and $\angle E$ as internal angles of the pentagon $A B C D E$. Let the perpendicular bisectors of $A C$ and $B D$, which pass respectively through $B$ and $C$, meet at point $I$. Then $B D \perp C I$ and, similarly, $A C \perp B I$. Hence $A C$ and $B D$ meet at the orthocenter $H$ of the triangle $B I C$, and $I H \perp B C$. It remains to prove that $E$ lies on the line $I H$ or, equivalently, $E I \perp B C$.

Lines $I B$ and $I C$ bisect $\angle B$ and $\angle C$, respectively. Since $I A=I C, I B=I D$, and $A B=$ $B C=C D$, the triangles $I A B, I C B$ and $I C D$ are congruent. Hence $\angle I A B=\angle I C B=$ $\angle C / 2=\angle A / 2$, so the line $I A$ bisects $\angle A$. Similarly, the line $I D$ bisects $\angle D$. Finally, the line $I E$ bisects $\angle E$ because $I$ lies on all the other four internal bisectors of the angles of the pentagon.

The sum of the internal angles in a pentagon is $540^{\circ}$, so

$$
\angle E=540^{\circ}-2 \angle A+2 \angle B
$$

In quadrilateral $A B I E$,

$$
\begin{aligned}
\angle B I E & =360^{\circ}-\angle E A B-\angle A B I-\angle A E I=360^{\circ}-\angle A-\frac{1}{2} \angle B-\frac{1}{2} \angle E \\
& =360^{\circ}-\angle A-\frac{1}{2} \angle B-\left(270^{\circ}-\angle A-\angle B\right) \\
& =90^{\circ}+\frac{1}{2} \angle B=90^{\circ}+\angle I B C,
\end{aligned}
$$

which means that $E I \perp B C$, completing the proof.


Solution 2. We present another proof of the fact that $E$ lies on line $I H$. Since all five internal bisectors of $A B C D E$ meet at $I$, this pentagon has an inscribed circle with center $I$. Let this circle touch side $B C$ at $T$.

Applying Brianchon's theorem to the (degenerate) hexagon $A B T C D E$ we conclude that $A C, B D$ and $E T$ are concurrent, so point $E$ also lies on line $I H T$, completing the proof.

Solution 3. We present yet another proof that $E I \perp B C$. In pentagon $A B C D E, \angle E<$ $180^{\circ} \Longleftrightarrow \angle A+\angle B+\angle C+\angle D>360^{\circ}$. Then $\angle A+\angle B=\angle C+\angle D>180^{\circ}$, so rays $E A$ and $C B$ meet at a point $P$, and rays $B C$ and $E D$ meet at a point $Q$. Now,

$$
\angle P B A=180^{\circ}-\angle B=180^{\circ}-\angle D=\angle Q D C
$$

and, similarly, $\angle P A B=\angle Q C D$. Since $A B=C D$, the triangles $P A B$ and $Q C D$ are congruent with the same orientation. Moreover, $P Q E$ is isosceles with $E P=E Q$.


In Solution 1 we have proved that triangles $I A B$ and $I C D$ are also congruent with the same orientation. Then we conclude that quadrilaterals $P B I A$ and $Q D I C$ are congruent, which implies $I P=I Q$. Then $E I$ is the perpendicular bisector of $P Q$ and, therefore, $E I \perp$ $P Q \Longleftrightarrow E I \perp B C$.

Comment. Even though all three solutions used the point $I$, there are solutions that do not need it. We present an outline of such a solution: if $J$ is the incenter of $\triangle Q C D$ (with $P$ and $Q$ as defined in Solution 3), then a simple angle chasing shows that triangles $C J D$ and $B H C$ are congruent. Then if $S$ is the projection of $J$ onto side $C D$ and $T$ is the orthogonal projection of $H$ onto side $B C$, one can verify that

$$
Q T=Q C+C T=Q C+D S=Q C+\frac{C D+D Q-Q C}{2}=\frac{P B+B C+Q C}{2}=\frac{P Q}{2},
$$

so $T$ is the midpoint of $P Q$, and $E, H$ and $T$ all lie on the perpendicular bisector of $P Q$.

G2. Let $R$ and $S$ be distinct points on circle $\Omega$, and let $t$ denote the tangent line to $\Omega$ at $R$. Point $R^{\prime}$ is the reflection of $R$ with respect to $S$. A point $I$ is chosen on the smaller arc $R S$ of $\Omega$ so that the circumcircle $\Gamma$ of triangle $I S R^{\prime}$ intersects $t$ at two different points. Denote by $A$ the common point of $\Gamma$ and $t$ that is closest to $R$. Line $A I$ meets $\Omega$ again at $J$. Show that $J R^{\prime}$ is tangent to $\Gamma$.
(Luxembourg)
Solution 1. In the circles $\Omega$ and $\Gamma$ we have $\angle J R S=\angle J I S=\angle A R^{\prime} S$. On the other hand, since $R A$ is tangent to $\Omega$, we get $\angle S J R=\angle S R A$. So the triangles $A R R^{\prime}$ and $S J R$ are similar, and

$$
\frac{R^{\prime} R}{R J}=\frac{A R^{\prime}}{S R}=\frac{A R^{\prime}}{S R^{\prime}}
$$

The last relation, together with $\angle A R^{\prime} S=\angle J R R^{\prime}$, yields $\triangle A S R^{\prime} \sim \triangle R^{\prime} J R$, hence $\angle S A R^{\prime}=\angle R R^{\prime} J$. It follows that $J R^{\prime}$ is tangent to $\Gamma$ at $R^{\prime}$.


Solution 2. As in Solution 1, we notice that $\angle J R S=\angle J I S=\angle A R^{\prime} S$, so we have $R J \| A R^{\prime}$. Let $A^{\prime}$ be the reflection of $A$ about $S$; then $A R A^{\prime} R^{\prime}$ is a parallelogram with center $S$, and hence the point $J$ lies on the line $R A^{\prime}$.

From $\angle S R^{\prime} A^{\prime}=\angle S R A=\angle S J R$ we get that the points $S, J, A^{\prime}, R^{\prime}$ are concyclic. This proves that $\angle S R^{\prime} J=\angle S A^{\prime} J=\angle S A^{\prime} R=\angle S A R^{\prime}$, so $J R^{\prime}$ is tangent to $\Gamma$ at $R^{\prime}$.

G3. Let $O$ be the circumcenter of an acute scalene triangle $A B C$. Line $O A$ intersects the altitudes of $A B C$ through $B$ and $C$ at $P$ and $Q$, respectively. The altitudes meet at $H$. Prove that the circumcenter of triangle $P Q H$ lies on a median of triangle $A B C$.
(Ukraine)
Solution. Suppose, without loss of generality, that $A B<A C$. We have $\angle P Q H=90^{\circ}-$ $\angle Q A B=90^{\circ}-\angle O A B=\frac{1}{2} \angle A O B=\angle A C B$, and similarly $\angle Q P H=\angle A B C$. Thus triangles $A B C$ and $H P Q$ are similar. Let $\Omega$ and $\omega$ be the circumcircles of $A B C$ and $H P Q$, respectively. Since $\angle A H P=90^{\circ}-\angle H A C=\angle A C B=\angle H Q P$, line $A H$ is tangent to $\omega$.


Let $T$ be the center of $\omega$ and let lines $A T$ and $B C$ meet at $M$. We will take advantage of the similarity between $A B C$ and $H P Q$ and the fact that $A H$ is tangent to $\omega$ at $H$, with $A$ on line $P Q$. Consider the corresponding tangent $A S$ to $\Omega$, with $S \in B C$. Then $S$ and $A$ correspond to each other in $\triangle A B C \sim \triangle H P Q$, and therefore $\angle O S M=\angle O A T=\angle O A M$. Hence quadrilateral $S A O M$ is cyclic, and since the tangent line $A S$ is perpendicular to $A O$, $\angle O M S=180^{\circ}-\angle O A S=90^{\circ}$. This means that $M$ is the orthogonal projection of $O$ onto $B C$, which is its midpoint. So $T$ lies on median $A M$ of triangle $A B C$.

G4. In triangle $A B C$, let $\omega$ be the excircle opposite $A$. Let $D, E$, and $F$ be the points where $\omega$ is tangent to lines $B C, C A$, and $A B$, respectively. The circle $A E F$ intersects line $B C$ at $P$ and $Q$. Let $M$ be the midpoint of $A D$. Prove that the circle $M P Q$ is tangent to $\omega$.
(Denmark)
Solution 1. Denote by $\Omega$ the circle $A E F P Q$, and denote by $\gamma$ the circle $P Q M$. Let the line $A D$ meet $\omega$ again at $T \neq D$. We will show that $\gamma$ is tangent to $\omega$ at $T$.

We first prove that points $P, Q, M, T$ are concyclic. Let $A^{\prime}$ be the center of $\omega$. Since $A^{\prime} E \perp A E$ and $A^{\prime} F \perp A F, A A^{\prime}$ is a diameter in $\Omega$. Let $N$ be the midpoint of $D T$; from $A^{\prime} D=A^{\prime} T$ we can see that $\angle A^{\prime} N A=90^{\circ}$ and therefore $N$ also lies on the circle $\Omega$. Now, from the power of $D$ with respect to the circles $\gamma$ and $\Omega$ we get

$$
D P \cdot D Q=D A \cdot D N=2 D M \cdot \frac{D T}{2}=D M \cdot D T
$$

so $P, Q, M, T$ are concyclic.
If $E F \| B C$, then $A B C$ is isosceles and the problem is now immediate by symmetry. Otherwise, let the tangent line to $\omega$ at $T$ meet line $B C$ at point $R$. The tangent line segments $R D$ and $R T$ have the same length, so $A^{\prime} R$ is the perpendicular bisector of $D T$; since $N D=N T$, $N$ lies on this perpendicular bisector.

In right triangle $A^{\prime} R D, R D^{2}=R N \cdot R A^{\prime}=R P \cdot R Q$, in which the last equality was obtained from the power of $R$ with respect to $\Omega$. Hence $R T^{2}=R P \cdot R Q$, which implies that $R T$ is also tangent to $\gamma$. Because $R T$ is a common tangent to $\omega$ and $\gamma$, these two circles are tangent at $T$.


Solution 2. After proving that $P, Q, M, T$ are concyclic, we finish the problem in a different fashion. We only consider the case in which $E F$ and $B C$ are not parallel. Let lines $P Q$ and $E F$ meet at point $R$. Since $P Q$ and $E F$ are radical axes of $\Omega, \gamma$ and $\omega, \gamma$, respectively, $R$ is the radical center of these three circles.

With respect to the circle $\omega$, the line $D R$ is the polar of $D$, and the line $E F$ is the polar of $A$. So the pole of line $A D T$ is $D R \cap E F=R$, and therefore $R T$ is tangent to $\omega$.

Finally, since $T$ belongs to $\gamma$ and $\omega$ and $R$ is the radical center of $\gamma, \omega$ and $\Omega$, line $R T$ is the radical axis of $\gamma$ and $\omega$, and since it is tangent to $\omega$, it is also tangent to $\gamma$. Because $R T$ is a common tangent to $\omega$ and $\gamma$, these two circles are tangent at $T$.

Comment. In Solution 2 we defined the point $R$ from Solution 1 in a different way.

Solution 3. We give an alternative proof that the circles are tangent at the common point $T$. Again, we start from the fact that $P, Q, M, T$ are concyclic. Let point $O$ be the midpoint of diameter $A A^{\prime}$. Then $M O$ is the midline of triangle $A D A^{\prime}$, so $M O \| A^{\prime} D$. Since $A^{\prime} D \perp P Q$, $M O$ is perpendicular to $P Q$ as well.

Looking at circle $\Omega$, which has center $O, M O \perp P Q$ implies that $M O$ is the perpendicular bisector of the chord $P Q$. Thus $M$ is the midpoint of arc $\widehat{P Q}$ from $\gamma$, and the tangent line $m$ to $\gamma$ at $M$ is parallel to $P Q$.


Consider the homothety with center $T$ and ratio $\frac{T D}{T M}$. It takes $D$ to $M$, and the line $P Q$ to the line $m$. Since the circle that is tangent to a line at a given point and that goes through another given point is unique, this homothety also takes $\omega$ (tangent to $P Q$ and going through $T$ ) to $\gamma$ (tangent to $m$ and going through $T$ ). We conclude that $\omega$ and $\gamma$ are tangent at $T$.

G5. Let $A B C C_{1} B_{1} A_{1}$ be a convex hexagon such that $A B=B C$, and suppose that the line segments $A A_{1}, B B_{1}$, and $C C_{1}$ have the same perpendicular bisector. Let the diagonals $A C_{1}$ and $A_{1} C$ meet at $D$, and denote by $\omega$ the circle $A B C$. Let $\omega$ intersect the circle $A_{1} B C_{1}$ again at $E \neq B$. Prove that the lines $B B_{1}$ and $D E$ intersect on $\omega$.
(Ukraine)
Solution 1. If $A A_{1}=C C_{1}$, then the hexagon is symmetric about the line $B B_{1}$; in particular the circles $A B C$ and $A_{1} B C_{1}$ are tangent to each other. So $A A_{1}$ and $C C_{1}$ must be different. Since the points $A$ and $A_{1}$ can be interchanged with $C$ and $C_{1}$, respectively, we may assume $A A_{1}<C C_{1}$.

Let $R$ be the radical center of the circles $A E B C$ and $A_{1} E B C_{1}$, and the circumcircle of the symmetric trapezoid $A C C_{1} A_{1}$; that is the common point of the pairwise radical axes $A C, A_{1} C_{1}$, and $B E$. By the symmetry of $A C$ and $A_{1} C_{1}$, the point $R$ lies on the common perpendicular bisector of $A A_{1}$ and $C C_{1}$, which is the external bisector of $\angle A D C$.

Let $F$ be the second intersection of the line $D R$ and the circle $A C D$. From the power of $R$ with respect to the circles $\omega$ and $A C F D$ we have $R B \cdot R E=R A \cdot R C=R D \cdot D F$, so the points $B, E, D$ and $F$ are concyclic.

The line $R D F$ is the external bisector of $\angle A D C$, so the point $F$ bisects the arc $\overline{C D A}$. By $A B=B C$, on circle $\omega$, the point $B$ is the midpoint of arc $\overline{A E C}$; let $M$ be the point diametrically opposite to $B$, that is the midpoint of the opposite $\operatorname{arc} \widetilde{C A}$ of $\omega$. Notice that the points $B, F$ and $M$ lie on the perpendicular bisector of $A C$, so they are collinear.


Finally, let $X$ be the second intersection point of $\omega$ and the line $D E$. Since $B M$ is a diameter in $\omega$, we have $\angle B X M=90^{\circ}$. Moreover,

$$
\angle E X M=180^{\circ}-\angle M B E=180^{\circ}-\angle F B E=\angle E D F,
$$

so $M X$ and $F D$ are parallel. Since $B X$ is perpendicular to $M X$ and $B B_{1}$ is perpendicular to $F D$, this shows that $X$ lies on line $B B_{1}$.

Solution 2. Define point $M$ as the point opposite to $B$ on circle $\omega$, and point $R$ as the intersection of lines $A C, A_{1} C_{1}$ and $B E$, and show that $R$ lies on the external bisector of $\angle A D C$, like in the first solution.

Since $B$ is the midpoint of the arc $\widehat{A E C}$, the line $B E R$ is the external bisector of $\angle C E A$. Now we show that the internal angle bisectors of $\angle A D C$ and $\angle C E A$ meet on the segment $A C$. Let the angle bisector of $\angle A D C$ meet $A C$ at $S$, and let the angle bisector of $\angle C E A$, which is line $E M$, meet $A C$ at $S^{\prime}$. By applying the angle bisector theorem to both internal and external bisectors of $\angle A D C$ and $\angle C E A$,

$$
A S: C S=A D: C D=A R: C R=A E: C E=A S^{\prime}: C S^{\prime}
$$

so indeed $S=S^{\prime}$.
By $\angle R D S=\angle S E R=90^{\circ}$ the points $R, S, D$ and $E$ are concyclic.


Now let the lines $B B_{1}$ and $D E$ meet at point $X$. Notice that $\angle E X B=\angle E D S$ because both $B B_{1}$ and $D S$ are perpendicular to the line $D R$, we have that $\angle E D S=\angle E R S$ in circle $S R D E$, and $\angle E R S=\angle E M B$ because $S R \perp B M$ and $E R \perp M E$. Therefore, $\angle E X B=\angle E M B$, so indeed, the point $X$ lies on $\omega$.

G6. Let $n \geqslant 3$ be an integer. Two regular $n$-gons $\mathcal{A}$ and $\mathcal{B}$ are given in the plane. Prove that the vertices of $\mathcal{A}$ that lie inside $\mathcal{B}$ or on its boundary are consecutive.
(That is, prove that there exists a line separating those vertices of $\mathcal{A}$ that lie inside $\mathcal{B}$ or on its boundary from the other vertices of $\mathcal{A}$.)
(Czech Republic)
Solution 1. In both solutions, by a polygon we always mean its interior together with its boundary.

We start with finding a regular $n$-gon $\mathcal{C}$ which $(i)$ is inscribed into $\mathcal{B}$ (that is, all vertices of $\mathcal{C}$ lie on the perimeter of $\mathcal{B}$ ); and (ii) is either a translation of $\mathcal{A}$, or a homothetic image of $\mathcal{A}$ with a positive factor.

Such a polygon may be constructed as follows. Let $O_{A}$ and $O_{B}$ be the centers of $\mathcal{A}$ and $\mathcal{B}$, respectively, and let $A$ be an arbitrary vertex of $\mathcal{A}$. Let $\overrightarrow{O_{B} C}$ be the vector co-directional to $\overrightarrow{O_{A} A}$, with $C$ lying on the perimeter of $\mathcal{B}$. The rotations of $C$ around $O_{B}$ by multiples of $2 \pi / n$ form the required polygon. Indeed, it is regular, inscribed into $\mathcal{B}$ (due to the rotational symmetry of $\mathcal{B}$ ), and finally the translation/homothety mapping $\overrightarrow{O_{A} A}$ to $\overrightarrow{O_{B} C}$ maps $\mathcal{A}$ to $\mathcal{C}$.

Now we separate two cases.


Construction of $\mathcal{C}$


Case 1: Translation

Case 1: $\mathcal{C}$ is a translation of $\mathcal{A}$ by a vector $\vec{v}$.
Denote by $t$ the translation transform by vector $\vec{v}$. We need to prove that the vertices of $\mathcal{C}$ which stay in $\mathcal{B}$ under $t$ are consecutive. To visualize the argument, we refer the plane to Cartesian coordinates so that the $x$-axis is co-directional with $\vec{v}$. This way, the notions of right/left and top/bottom are also introduced, according to the $x$ - and $y$-coordinates, respectively.

Let $B_{\mathrm{T}}$ and $B_{\mathrm{B}}$ be the top and the bottom vertices of $\mathcal{B}$ (if several vertices are extremal, we take the rightmost of them). They split the perimeter of $\mathcal{B}$ into the right part $\mathcal{B}_{\mathrm{R}}$ and the left part $\mathcal{B}_{\mathrm{L}}$ (the vertices $B_{\mathrm{T}}$ and $B_{\mathrm{B}}$ are assumed to lie in both parts); each part forms a connected subset of the perimeter of $\mathcal{B}$. So the vertices of $\mathcal{C}$ are also split into two parts $\mathcal{C}_{\mathrm{L}} \subset \mathcal{B}_{\mathrm{L}}$ and $\mathcal{C}_{\mathrm{R}} \subset \mathcal{B}_{\mathrm{R}}$, each of which consists of consecutive vertices.

Now, all the points in $\mathcal{B}_{\mathrm{R}}$ (and hence in $\mathcal{C}_{\mathrm{R}}$ ) move out from $\mathcal{B}$ under $t$, since they are the rightmost points of $\mathcal{B}$ on the corresponding horizontal lines. It remains to prove that the vertices of $\mathcal{C}_{\mathrm{L}}$ which stay in $\mathcal{B}$ under $t$ are consecutive.

For this purpose, let $C_{1}, C_{2}$, and $C_{3}$ be three vertices in $\mathcal{C}_{\mathrm{L}}$ such that $C_{2}$ is between $C_{1}$ and $C_{3}$, and $t\left(C_{1}\right)$ and $t\left(C_{3}\right)$ lie in $\mathcal{B}$; we need to prove that $t\left(C_{2}\right) \in \mathcal{B}$ as well. Let $A_{i}=t\left(C_{i}\right)$. The line through $C_{2}$ parallel to $\vec{v}$ crosses the segment $C_{1} C_{3}$ to the right of $C_{2}$; this means that this line crosses $A_{1} A_{3}$ to the right of $A_{2}$, so $A_{2}$ lies inside the triangle $A_{1} C_{2} A_{3}$ which is contained in $\mathcal{B}$. This yields the desired result.

Case 2: $\mathcal{C}$ is a homothetic image of $\mathcal{A}$ centered at $X$ with factor $k>0$.

Denote by $h$ the homothety mapping $\mathcal{C}$ to $\mathcal{A}$. We need now to prove that the vertices of $\mathcal{C}$ which stay in $\mathcal{B}$ after applying $h$ are consecutive. If $X \in \mathcal{B}$, the claim is easy. Indeed, if $k<1$, then the vertices of $\mathcal{A}$ lie on the segments of the form $X C(C$ being a vertex of $\mathcal{C})$ which lie in $\mathcal{B}$. If $k>1$, then the vertices of $\mathcal{A}$ lie on the extensions of such segments $X C$ beyond $C$, and almost all these extensions lie outside $\mathcal{B}$. The exceptions may occur only in case when $X$ lies on the boundary of $\mathcal{B}$, and they may cause one or two vertices of $\mathcal{A}$ stay on the boundary of $\mathcal{B}$. But even in this case those vertices are still consecutive.

So, from now on we assume that $X \notin \mathcal{B}$.
Now, there are two vertices $B_{\mathrm{T}}$ and $\mathcal{B}_{\mathrm{B}}$ of $\mathcal{B}$ such that $\mathcal{B}$ is contained in the angle $\angle B_{\mathrm{T}} X B_{\mathrm{B}}$; if there are several options, say, for $B_{\mathrm{T}}$, then we choose the farthest one from $X$ if $k>1$, and the nearest one if $k<1$. For the visualization purposes, we refer the plane to Cartesian coordinates so that the $y$-axis is co-directional with $\overrightarrow{B_{\mathrm{B}} B_{\mathrm{T}}}$, and $X$ lies to the left of the line $B_{\mathrm{T}} B_{\mathrm{B}}$. Again, the perimeter of $\mathcal{B}$ is split by $B_{\mathrm{T}}$ and $B_{\mathrm{B}}$ into the right part $\mathcal{B}_{\mathrm{R}}$ and the left part $\mathcal{B}_{\mathrm{L}}$, and the set of vertices of $\mathcal{C}$ is split into two subsets $\mathcal{C}_{\mathrm{R}} \subset \mathcal{B}_{\mathrm{R}}$ and $\mathcal{C}_{\mathrm{L}} \subset \mathcal{B}_{\mathrm{L}}$.


Case 2, $X$ inside $\mathcal{B}$


Subcase 2.1: $k>1$

Subcase 2.1: $k>1$.
In this subcase, all points from $\mathcal{B}_{\mathrm{R}}$ (and hence from $\mathcal{C}_{\mathrm{R}}$ ) move out from $\mathcal{B}$ under $h$, because they are the farthest points of $\mathcal{B}$ on the corresponding rays emanated from $X$. It remains to prove that the vertices of $\mathcal{C}_{\mathrm{L}}$ which stay in $\mathcal{B}$ under $h$ are consecutive.

Again, let $C_{1}, C_{2}, C_{3}$ be three vertices in $\mathcal{C}_{\mathrm{L}}$ such that $C_{2}$ is between $C_{1}$ and $C_{3}$, and $h\left(C_{1}\right)$ and $h\left(C_{3}\right)$ lie in $\mathcal{B}$. Let $A_{i}=h\left(C_{i}\right)$. Then the ray $X C_{2}$ crosses the segment $C_{1} C_{3}$ beyond $C_{2}$, so this ray crosses $A_{1} A_{3}$ beyond $A_{2}$; this implies that $A_{2}$ lies in the triangle $A_{1} C_{2} A_{3}$, which is contained in $\mathcal{B}$.


Subcase 2.2: $k<1$
Subcase 2.2: $k<1$.
This case is completely similar to the previous one. All points from $\mathcal{B}_{\mathrm{L}}$ (and hence from $\mathcal{C}_{\mathrm{L}}$ move out from $\mathcal{B}$ under $h$, because they are the nearest points of $\mathcal{B}$ on the corresponding
rays emanated from $X$. Assume that $C_{1}, C_{2}$, and $C_{3}$ are three vertices in $\mathcal{C}_{\mathrm{R}}$ such that $C_{2}$ lies between $C_{1}$ and $C_{3}$, and $h\left(C_{1}\right)$ and $h\left(C_{3}\right)$ lie in $\mathcal{B}$; let $A_{i}=h\left(C_{i}\right)$. Then $A_{2}$ lies on the segment $X C_{2}$, and the segments $X A_{2}$ and $A_{1} A_{3}$ cross each other. Thus $A_{2}$ lies in the triangle $A_{1} C_{2} A_{3}$, which is contained in $\mathcal{B}$.

Comment 1. In fact, Case 1 can be reduced to Case 2 via the following argument.
Assume that $\mathcal{A}$ and $\mathcal{C}$ are congruent. Apply to $\mathcal{A}$ a homothety centered at $O_{B}$ with a factor slightly smaller than 1 to obtain a polygon $\mathcal{A}^{\prime}$. With appropriately chosen factor, the vertices of $\mathcal{A}$ which were outside $/$ inside $\mathcal{B}$ stay outside/inside it, so it suffices to prove our claim for $\mathcal{A}^{\prime}$ instead of $\mathcal{A}$. And now, the polygon $\mathcal{A}^{\prime}$ is a homothetic image of $\mathcal{C}$, so the arguments from Case 2 apply.

Comment 2. After the polygon $\mathcal{C}$ has been found, the rest of the solution uses only the convexity of the polygons, instead of regularity. Thus, it proves a more general statement:

Assume that $\mathcal{A}, \mathcal{B}$, and $\mathcal{C}$ are three convex polygons in the plane such that $\mathcal{C}$ is inscribed into $\mathcal{B}$, and $\mathcal{A}$ can be obtained from it via either translation or positive homothety. Then the vertices of $\mathcal{A}$ that lie inside $\mathcal{B}$ or on its boundary are consecutive.
Solution 2. Let $O_{A}$ and $O_{B}$ be the centers of $\mathcal{A}$ and $\mathcal{B}$, respectively. Denote $[n]=\{1,2, \ldots, n\}$.
We start with introducing appropriate enumerations and notations. Enumerate the sidelines of $\mathcal{B}$ clockwise as $\ell_{1}, \ell_{2}, \ldots, \ell_{n}$. Denote by $\mathcal{H}_{i}$ the half-plane of $\ell_{i}$ that contains $\mathcal{B}\left(\mathcal{H}_{i}\right.$ is assumed to contain $\ell_{i}$ ); by $B_{i}$ the midpoint of the side belonging to $\ell_{i}$; and finally denote $\overrightarrow{b_{i}}=\overrightarrow{B_{i} O_{B}}$. (As usual, the numbering is cyclic modulo $n$, so $\ell_{n+i}=\ell_{i}$ etc.)

Now, choose a vertex $A_{1}$ of $\mathcal{A}$ such that the vector $\overrightarrow{O_{A} A_{1}}$ points "mostly outside $\mathcal{H}_{1}$ "; strictly speaking, this means that the scalar product $\left\langle\overrightarrow{O_{A} A_{1}}, \overrightarrow{b_{1}}\right\rangle$ is minimal. Starting from $A_{1}$, enumerate the vertices of $\mathcal{A}$ clockwise as $A_{1}, A_{2}, \ldots, A_{n}$; by the rotational symmetry, the choice of $A_{1}$ yields that the vector $\overrightarrow{O_{A} A_{i}}$ points "mostly outside $\mathcal{H}_{i}$ ", i.e.,

$$
\begin{equation*}
\left\langle\overrightarrow{O_{A} A_{i}}, \overrightarrow{b_{i}}\right\rangle=\min _{j \in[n]}\left\langle\overrightarrow{O_{A} A_{j}}, \overrightarrow{b_{i}}\right\rangle . \tag{1}
\end{equation*}
$$



Enumerations and notations
We intend to reformulate the problem in more combinatorial terms, for which purpose we introduce the following notion. Say that a subset $I \subseteq[n]$ is connected if the elements of this set are consecutive in the cyclic order (in other words, if we join each $i$ with $i+1 \bmod n$ by an edge, this subset is connected in the usual graph sense). Clearly, the union of two connected subsets sharing at least one element is connected too. Next, for any half-plane $\mathcal{H}$ the indices of vertices of, say, $\mathcal{A}$ that lie in $\mathcal{H}$ form a connected set.

To access the problem, we denote

$$
M=\left\{j \in[n]: A_{j} \notin \mathcal{B}\right\}, \quad M_{i}=\left\{j \in[n]: A_{j} \notin \mathcal{H}_{i}\right\} \quad \text { for } i \in[n] .
$$

We need to prove that $[n] \backslash M$ is connected, which is equivalent to $M$ being connected. On the other hand, since $\mathcal{B}=\bigcap_{i \in[n]} \mathcal{H}_{i}$, we have $M=\bigcup_{i \in[n]} M_{i}$, where the sets $M_{i}$ are easier to investigate. We will utilize the following properties of these sets; the first one holds by the definition of $M_{i}$, along with the above remark.


The sets $M_{i}$

Property 1: Each set $M_{i}$ is connected.
Property 2: If $M_{i}$ is nonempty, then $i \in M_{i}$.
Proof. Indeed, we have

$$
\begin{equation*}
j \in M_{i} \Longleftrightarrow A_{j} \notin \mathcal{H}_{i} \Longleftrightarrow\left\langle\overrightarrow{B_{i} A_{j}}, \overrightarrow{b_{i}}\right\rangle<0 \Longleftrightarrow\left\langle\overrightarrow{O_{A} A_{j}}, \overrightarrow{b_{i}}\right\rangle<\left\langle\overrightarrow{O_{A} B_{i}}, \overrightarrow{b_{i}}\right\rangle \tag{2}
\end{equation*}
$$

The right-hand part of the last inequality does not depend on $j$. Therefore, if some $j$ lies in $M_{i}$, then by (1) so does $i$.

In view of Property 2 , it is useful to define the set

$$
M^{\prime}=\left\{i \in[n]: i \in M_{i}\right\}=\left\{i \in[n]: M_{i} \neq \varnothing\right\} .
$$

Property 3: The set $M^{\prime}$ is connected.
Proof. To prove this property, we proceed on with the investigation started in (2) to write

$$
i \in M^{\prime} \Longleftrightarrow A_{i} \in M_{i} \Longleftrightarrow\left\langle\overrightarrow{B_{i} A_{i}}, \overrightarrow{b_{i}}\right\rangle<0 \Longleftrightarrow\left\langle\overrightarrow{O_{B} O_{A}}, \overrightarrow{b_{i}}\right\rangle<\left\langle\overrightarrow{O_{B} B_{i}}, \overrightarrow{b_{i}}\right\rangle+\left\langle\overrightarrow{A_{i} O_{A}}, \overrightarrow{b_{i}}\right\rangle .
$$

The right-hand part of the obtained inequality does not depend on $i$, due to the rotational symmetry; denote its constant value by $\mu$. Thus, $i \in M^{\prime}$ if and only if $\left\langle\overrightarrow{O_{B} O_{A}}, \overrightarrow{b_{i}}\right\rangle<\mu$. This condition is in turn equivalent to the fact that $B_{i}$ lies in a certain (open) half-plane whose boundary line is orthogonal to $O_{B} O_{A}$; thus, it defines a connected set.

Now we can finish the solution. Since $M^{\prime} \subseteq M$, we have

$$
M=\bigcup_{i \in[n]} M_{i}=M^{\prime} \cup \bigcup_{i \in[n]} M_{i},
$$

so $M$ can be obtained from $M^{\prime}$ by adding all the sets $M_{i}$ one by one. All these sets are connected, and each nonempty $M_{i}$ contains an element of $M^{\prime}$ (namely, $i$ ). Thus their union is also connected.

Comment 3. Here we present a way in which one can come up with a solution like the one above.
Assume, for sake of simplicity, that $O_{A}$ lies inside $\mathcal{B}$. Let us first put onto the plane a very small regular $n$-gon $\mathcal{A}^{\prime}$ centered at $O_{A}$ and aligned with $\mathcal{A}$; all its vertices lie inside $\mathcal{B}$. Now we start blowing it up, looking at the order in which the vertices leave $\mathcal{B}$. To go out of $\mathcal{B}$, a vertex should cross a certain side of $\mathcal{B}$ (which is hard to describe), or, equivalently, to cross at least one sideline of $\mathcal{B}$ - and this event is easier to describe. Indeed, the first vertex of $\mathcal{A}^{\prime}$ to cross $\ell_{i}$ is the vertex $A_{i}^{\prime}$ (corresponding to $A_{i}$ in $\mathcal{A}$ ); more generally, the vertices $A_{j}^{\prime}$ cross $\ell_{i}$ in such an order that the scalar product $\left\langle\overrightarrow{O_{A} \overrightarrow{A_{j}}}, \overrightarrow{b_{i}}\right\rangle$ does not increase. For different indices $i$, these orders are just cyclic shifts of each other; and this provides some intuition for the notions and claims from Solution 2.

G7. A convex quadrilateral $A B C D$ has an inscribed circle with center $I$. Let $I_{a}, I_{b}, I_{c}$, and $I_{d}$ be the incenters of the triangles $D A B, A B C, B C D$, and $C D A$, respectively. Suppose that the common external tangents of the circles $A I_{b} I_{d}$ and $C I_{b} I_{d}$ meet at $X$, and the common external tangents of the circles $B I_{a} I_{c}$ and $D I_{a} I_{c}$ meet at $Y$. Prove that $\angle X I Y=90^{\circ}$.
(Kazakhstan)
Solution. Denote by $\omega_{a}, \omega_{b}, \omega_{c}$ and $\omega_{d}$ the circles $A I_{b} I_{d}, B I_{a} I_{c}, C I_{b} I_{d}$, and $D I_{a} I_{c}$, let their centers be $O_{a}, O_{b}, O_{c}$ and $O_{d}$, and let their radii be $r_{a}, r_{b}, r_{c}$ and $r_{d}$, respectively.
Claim 1. $I_{b} I_{d} \perp A C$ and $I_{a} I_{c} \perp B D$.
Proof. Let the incircles of triangles $A B C$ and $A C D$ be tangent to the line $A C$ at $T$ and $T^{\prime}$, respectively. (See the figure to the left.) We have $A T=\frac{A B+A C-B C}{2}$ in triangle $A B C, A T^{\prime}=$ $\frac{A D+A C-C D}{2}$ in triangle $A C D$, and $A B-B C=A D-C D$ in quadrilateral $A B C D$, so

$$
A T=\frac{A C+A B-B C}{2}=\frac{A C+A D-C D}{2}=A T^{\prime}
$$

This shows $T=T^{\prime}$. As an immediate consequence, $I_{b} I_{d} \perp A C$.
The second statement can be shown analogously.


Claim 2. The points $O_{a}, O_{b}, O_{c}$ and $O_{d}$ lie on the lines $A I, B I, C I$ and $D I$, respectively. Proof. By symmetry it suffices to prove the claim for $O_{a}$. (See the figure to the right above.)

Notice first that the incircles of triangles $A B C$ and $A C D$ can be obtained from the incircle of the quadrilateral $A B C D$ with homothety centers $B$ and $D$, respectively, and homothety factors less than 1 , therefore the points $I_{b}$ and $I_{d}$ lie on the line segments $B I$ and $D I$, respectively.

As is well-known, in every triangle the altitude and the diameter of the circumcircle starting from the same vertex are symmetric about the angle bisector. By Claim 1, in triangle $A I_{d} I_{b}$, the segment $A T$ is the altitude starting from $A$. Since the foot $T$ lies inside the segment $I_{b} I_{d}$, the circumcenter $O_{a}$ of triangle $A I_{d} I_{b}$ lies in the angle domain $I_{b} A I_{d}$ in such a way that $\angle I_{b} A T=\angle O_{a} A I_{d}$. The points $I_{b}$ and $I_{d}$ are the incenters of triangles $A B C$ and $A C D$, so the lines $A I_{b}$ and $A I_{d}$ bisect the angles $\angle B A C$ and $\angle C A D$, respectively. Then

$$
\angle O_{a} A D=\angle O_{a} A I_{d}+\angle I_{d} A D=\angle I_{b} A T+\angle I_{d} A D=\frac{1}{2} \angle B A C+\frac{1}{2} \angle C A D=\frac{1}{2} \angle B A D,
$$

so $O_{a}$ lies on the angle bisector of $\angle B A D$, that is, on line $A I$.
The point $X$ is the external similitude center of $\omega_{a}$ and $\omega_{c}$; let $U$ be their internal similitude center. The points $O_{a}$ and $O_{c}$ lie on the perpendicular bisector of the common chord $I_{b} I_{d}$ of $\omega_{a}$ and $\omega_{c}$, and the two similitude centers $X$ and $U$ lie on the same line; by Claim 2, that line is parallel to $A C$.


From the similarity of the circles $\omega_{a}$ and $\omega_{c}$, from $O_{a} I_{b}=O_{a} I_{d}=O_{a} A=r_{a}$ and $O_{c} I_{b}=$ $O_{c} I_{d}=O_{c} C=r_{c}$, and from $A C \| O_{a} O_{c}$ we can see that

$$
\frac{O_{a} X}{O_{c} X}=\frac{O_{a} U}{O_{c} U}=\frac{r_{a}}{r_{c}}=\frac{O_{a} I_{b}}{O_{c} I_{b}}=\frac{O_{a} I_{d}}{O_{c} I_{d}}=\frac{O_{a} A}{O_{c} C}=\frac{O_{a} I}{O_{c} I} .
$$

So the points $X, U, I_{b}, I_{d}, I$ lie on the Apollonius circle of the points $O_{a}, O_{c}$ with ratio $r_{a}: r_{c}$. In this Apollonius circle $X U$ is a diameter, and the lines $I U$ and $I X$ are respectively the internal and external bisectors of $\angle O_{a} I O_{c}=\angle A I C$, according to the angle bisector theorem. Moreover, in the Apollonius circle the diameter $U X$ is the perpendicular bisector of $I_{b} I_{d}$, so the lines $I X$ and $I U$ are the internal and external bisectors of $\angle I_{b} I I_{d}=\angle B I D$, respectively.

Repeating the same argument for the points $B, D$ instead of $A, C$, we get that the line $I Y$ is the internal bisector of $\angle A I C$ and the external bisector of $\angle B I D$. Therefore, the lines $I X$ and $I Y$ respectively are the internal and external bisectors of $\angle B I D$, so they are perpendicular.

Comment. In fact the points $O_{a}, O_{b}, O_{c}$ and $O_{d}$ lie on the line segments $A I, B I, C I$ and $D I$, respectively. For the point $O_{a}$ this can be shown for example by $\angle I_{d} O_{a} A+\angle A O_{a} I_{b}=\left(180^{\circ}-\right.$ $\left.2 \angle O_{a} A I_{d}\right)+\left(180^{\circ}-2 \angle I_{b} A O_{a}\right)=360^{\circ}-\angle B A D=\angle A D I+\angle D I A+\angle A I B+\angle I B A>\angle I_{d} I A+\angle A I I_{b}$.

The solution also shows that the line $I Y$ passes through the point $U$, and analogously, $I X$ passes through the internal similitude center of $\omega_{b}$ and $\omega_{d}$.

G8. There are 2017 mutually external circles drawn on a blackboard, such that no two are tangent and no three share a common tangent. A tangent segment is a line segment that is a common tangent to two circles, starting at one tangent point and ending at the other one. Luciano is drawing tangent segments on the blackboard, one at a time, so that no tangent segment intersects any other circles or previously drawn tangent segments. Luciano keeps drawing tangent segments until no more can be drawn. Find all possible numbers of tangent segments when he stops drawing.
(Australia)
Answer: If there were $n$ circles, there would always be exactly $3(n-1)$ segments; so the only possible answer is $3 \cdot 2017-3=6048$.
Solution 1. First, consider a particular arrangement of circles $C_{1}, C_{2}, \ldots, C_{n}$ where all the centers are aligned and each $C_{i}$ is eclipsed from the other circles by its neighbors - for example, taking $C_{i}$ with center $\left(i^{2}, 0\right)$ and radius $i / 2$ works. Then the only tangent segments that can be drawn are between adjacent circles $C_{i}$ and $C_{i+1}$, and exactly three segments can be drawn for each pair. So Luciano will draw exactly $3(n-1)$ segments in this case.


For the general case, start from a final configuration (that is, an arrangement of circles and segments in which no further segments can be drawn). The idea of the solution is to continuously resize and move the circles around the plane, one by one (in particular, making sure we never have 4 circles with a common tangent line), and show that the number of segments drawn remains constant as the picture changes. This way, we can reduce any circle/segment configuration to the particular one mentioned above, and the final number of segments must remain at $3 n-3$.

Some preliminary considerations: look at all possible tangent segments joining any two circles. A segment that is tangent to a circle $A$ can do so in two possible orientations - it may come out of $A$ in clockwise or counterclockwise orientation. Two segments touching the same circle with the same orientation will never intersect each other. Each pair $(A, B)$ of circles has 4 choices of tangent segments, which can be identified by their orientations - for example, ( $A+, B-$ ) would be the segment which comes out of $A$ in clockwise orientation and comes out of $B$ in counterclockwise orientation. In total, we have $2 n(n-1)$ possible segments, disregarding intersections.

Now we pick a circle $C$ and start to continuously move and resize it, maintaining all existing tangent segments according to their identifications, including those involving $C$. We can keep our choice of tangent segments until the configuration reaches a transition. We lose nothing if we assume that $C$ is kept at least $\varepsilon$ units away from any other circle, where $\varepsilon$ is a positive, fixed constant; therefore at a transition either: (1) a currently drawn tangent segment $t$ suddenly becomes obstructed; or (2) a currently absent tangent segment $t$ suddenly becomes unobstructed and available.
Claim. A transition can only occur when three circles $C_{1}, C_{2}, C_{3}$ are tangent to a common line $\ell$ containing $t$, in a way such that the three tangent segments lying on $\ell$ (joining the three circles pairwise) are not obstructed by any other circles or tangent segments (other than $C_{1}, C_{2}, C_{3}$ ).
Proof. Since (2) is effectively the reverse of (1), it suffices to prove the claim for (1). Suppose $t$ has suddenly become obstructed, and let us consider two cases.

Case 1: $t$ becomes obstructed by a circle


Then the new circle becomes the third circle tangent to $\ell$, and no other circles or tangent segments are obstructing $t$.

## Case 2: $t$ becomes obstructed by another tangent segment $t^{\prime}$

When two segments $t$ and $t^{\prime}$ first intersect each other, they must do so at a vertex of one of them. But if a vertex of $t^{\prime}$ first crossed an interior point of $t$, the circle associated to this vertex was already blocking $t$ (absurd), or is about to (we already took care of this in case 1). So we only have to analyze the possibility of $t$ and $t^{\prime}$ suddenly having a common vertex. However, if that happens, this vertex must belong to a single circle (remember we are keeping different circles at least $\varepsilon$ units apart from each other throughout the moving/resizing process), and therefore they must have different orientations with respect to that circle.


Thus, at the transition moment, both $t$ and $t^{\prime}$ are tangent to the same circle at a common point, that is, they must be on the same line $\ell$ and hence we again have three circles simultaneously tangent to $\ell$. Also no other circles or tangent segments are obstructing $t$ or $t^{\prime}$ (otherwise, they would have disappeared before this transition).

Next, we focus on the maximality of a configuration immediately before and after a transition, where three circles share a common tangent line $\ell$. Let the three circles be $C_{1}, C_{2}, C_{3}$, ordered by their tangent points. The only possibly affected segments are the ones lying on $\ell$, namely $t_{12}, t_{23}$ and $t_{13}$. Since $C_{2}$ is in the middle, $t_{12}$ and $t_{23}$ must have different orientations with respect to $C_{2}$. For $C_{1}, t_{12}$ and $t_{13}$ must have the same orientation, while for $C_{3}, t_{13}$ and $t_{23}$ must have the same orientation. The figure below summarizes the situation, showing alternative positions for $C_{1}$ (namely, $C_{1}$ and $C_{1}^{\prime}$ ) and for $C_{3}\left(C_{3}\right.$ and $\left.C_{3}^{\prime}\right)$.


Now perturb the diagram slightly so the three circles no longer have a common tangent, while preserving the definition of $t_{12}, t_{23}$ and $t_{13}$ according to their identifications. First note that no other circles or tangent segments can obstruct any of these segments. Also recall that tangent segments joining the same circle at the same orientation will never obstruct each other.

The availability of the tangent segments can now be checked using simple diagrams.
Case 1: $t_{13}$ passes through $C_{2}$


In this case, $t_{13}$ is not available, but both $t_{12}$ and $t_{23}$ are.
Case 2: $t_{13}$ does not pass through $C_{2}$


Now $t_{13}$ is available, but $t_{12}$ and $t_{23}$ obstruct each other, so only one can be drawn.
In any case, exactly 2 out of these 3 segments can be drawn. Thus the maximal number of segments remains constant as we move or resize the circles, and we are done.

Solution 2. First note that all tangent segments lying on the boundary of the convex hull of the circles are always drawn since they do not intersect anything else. Now in the final picture, aside from the $n$ circles, the blackboard is divided into regions. We can consider the picture as a plane (multi-)graph $G$ in which the circles are the vertices and the tangent segments are the edges. The idea of this solution is to find a relation between the number of edges and the number of regions in $G$; then, once we prove that $G$ is connected, we can use Euler's formula to finish the problem.

The boundary of each region consists of 1 or more (for now) simple closed curves, each made of arcs and tangent segments. The segment and the arc might meet smoothly (as in $S_{i}$, $i=1,2, \ldots, 6$ in the figure below) or not (as in $P_{1}, P_{2}, P_{3}, P_{4}$; call such points sharp corners of the boundary). In other words, if a person walks along the border, her direction would suddenly turn an angle of $\pi$ at a sharp corner.


Claim 1. The outer boundary $B_{1}$ of any internal region has at least 3 sharp corners.
Proof. Let a person walk one lap along $B_{1}$ in the counterclockwise orientation. As she does so, she will turn clockwise as she moves along the circle arcs, and not turn at all when moving along the lines. On the other hand, her total rotation after one lap is $2 \pi$ in the counterclockwise direction! Where could she be turning counterclockwise? She can only do so at sharp corners, and, even then, she turns only an angle of $\pi$ there. But two sharp corners are not enough, since at least one arc must be present-so she must have gone through at least 3 sharp corners.

Claim 2. Each internal region is simply connected, that is, has only one boundary curve.
Proof. Suppose, by contradiction, that some region has an outer boundary $B_{1}$ and inner boundaries $B_{2}, B_{3}, \ldots, B_{m}(m \geqslant 2)$. Let $P_{1}$ be one of the sharp corners of $B_{1}$.

Now consider a car starting at $P_{1}$ and traveling counterclockwise along $B_{1}$. It starts in reverse, i.e., it is initially facing the corner $P_{1}$. Due to the tangent conditions, the car may travel in a way so that its orientation only changes when it is moving along an arc. In particular, this means the car will sometimes travel forward. For example, if the car approaches a sharp corner when driving in reverse, it would continue travel forward after the corner, instead of making an immediate half-turn. This way, the orientation of the car only changes in a clockwise direction since the car always travels clockwise around each arc.

Now imagine there is a laser pointer at the front of the car, pointing directly ahead. Initially, the laser endpoint hits $P_{1}$, but, as soon as the car hits an arc, the endpoint moves clockwise around $B_{1}$. In fact, the laser endpoint must move continuously along $B_{1}$ ! Indeed, if the endpoint ever jumped (within $B_{1}$, or from $B_{1}$ to one of the inner boundaries), at the moment of the jump the interrupted laser would be a drawable tangent segment that Luciano missed (see figure below for an example).


Now, let $P_{2}$ and $P_{3}$ be the next two sharp corners the car goes through, after $P_{1}$ (the previous lemma assures their existence). At $P_{2}$ the car starts moving forward, and at $P_{3}$ it will start to move in reverse again. So, at $P_{3}$, the laser endpoint is at $P_{3}$ itself. So while the car moved counterclockwise between $P_{1}$ and $P_{3}$, the laser endpoint moved clockwise between $P_{1}$ and $P_{3}$. That means the laser beam itself scanned the whole region within $B_{1}$, and it should have crossed some of the inner boundaries.

Claim 3. Each region has exactly 3 sharp corners.
Proof. Consider again the car of the previous claim, with its laser still firmly attached to its front, traveling the same way as before and going through the same consecutive sharp corners $P_{1}, P_{2}$ and $P_{3}$. As we have seen, as the car goes counterclockwise from $P_{1}$ to $P_{3}$, the laser endpoint goes clockwise from $P_{1}$ to $P_{3}$, so together they cover the whole boundary. If there were a fourth sharp corner $P_{4}$, at some moment the laser endpoint would pass through it. But, since $P_{4}$ is a sharp corner, this means the car must be on the extension of a tangent segment going through $P_{4}$. Since the car is not on that segment itself (the car never goes through $P_{4}$ ), we would have 3 circles with a common tangent line, which is not allowed.


We are now ready to finish the solution. Let $r$ be the number of internal regions, and $s$ be the number of tangent segments. Since each tangent segment contributes exactly 2 sharp corners to the diagram, and each region has exactly 3 sharp corners, we must have $2 s=3 r$. Since the graph corresponding to the diagram is connected, we can use Euler's formula $n-s+r=1$ and find $s=3 n-3$ and $r=2 n-2$.

## Number Theory

N1. The sequence $a_{0}, a_{1}, a_{2}, \ldots$ of positive integers satisfies

$$
a_{n+1}=\left\{\begin{array}{ll}
\sqrt{a_{n}}, & \text { if } \sqrt{a_{n}} \text { is an integer } \\
a_{n}+3, & \text { otherwise }
\end{array} \quad \text { for every } n \geqslant 0\right.
$$

Determine all values of $a_{0}>1$ for which there is at least one number $a$ such that $a_{n}=a$ for infinitely many values of $n$.
(South Africa)
Answer: All positive multiples of 3 .
Solution. Since the value of $a_{n+1}$ only depends on the value of $a_{n}$, if $a_{n}=a_{m}$ for two different indices $n$ and $m$, then the sequence is eventually periodic. So we look for the values of $a_{0}$ for which the sequence is eventually periodic.
Claim 1. If $a_{n} \equiv-1(\bmod 3)$, then, for all $m>n, a_{m}$ is not a perfect square. It follows that the sequence is eventually strictly increasing, so it is not eventually periodic.
Proof. A square cannot be congruent to -1 modulo 3 , so $a_{n} \equiv-1(\bmod 3)$ implies that $a_{n}$ is not a square, therefore $a_{n+1}=a_{n}+3>a_{n}$. As a consequence, $a_{n+1} \equiv a_{n} \equiv-1(\bmod 3)$, so $a_{n+1}$ is not a square either. By repeating the argument, we prove that, from $a_{n}$ on, all terms of the sequence are not perfect squares and are greater than their predecessors, which completes the proof.
Claim 2. If $a_{n} \not \equiv-1(\bmod 3)$ and $a_{n}>9$ then there is an index $m>n$ such that $a_{m}<a_{n}$.
Proof. Let $t^{2}$ be the largest perfect square which is less than $a_{n}$. Since $a_{n}>9, t$ is at least 3. The first square in the sequence $a_{n}, a_{n}+3, a_{n}+6, \ldots$ will be $(t+1)^{2},(t+2)^{2}$ or $(t+3)^{2}$, therefore there is an index $m>n$ such that $a_{m} \leqslant t+3<t^{2}<a_{n}$, as claimed.
Claim 3. If $a_{n} \equiv 0(\bmod 3)$, then there is an index $m>n$ such that $a_{m}=3$.
Proof. First we notice that, by the definition of the sequence, a multiple of 3 is always followed by another multiple of 3 . If $a_{n} \in\{3,6,9\}$ the sequence will eventually follow the periodic pattern $3,6,9,3,6,9, \ldots$. If $a_{n}>9$, let $j$ be an index such that $a_{j}$ is equal to the minimum value of the set $\left\{a_{n+1}, a_{n+2}, \ldots\right\}$. We must have $a_{j} \leqslant 9$, otherwise we could apply Claim 2 to $a_{j}$ and get a contradiction on the minimality hypothesis. It follows that $a_{j} \in\{3,6,9\}$, and the proof is complete.
Claim 4. If $a_{n} \equiv 1(\bmod 3)$, then there is an index $m>n$ such that $a_{m} \equiv-1(\bmod 3)$.
Proof. In the sequence, 4 is always followed by $2 \equiv-1(\bmod 3)$, so the claim is true for $a_{n}=4$. If $a_{n}=7$, the next terms will be $10,13,16,4,2, \ldots$ and the claim is also true. For $a_{n} \geqslant 10$, we again take an index $j>n$ such that $a_{j}$ is equal to the minimum value of the set $\left\{a_{n+1}, a_{n+2}, \ldots\right\}$, which by the definition of the sequence consists of non-multiples of 3 . Suppose $a_{j} \equiv 1(\bmod 3)$. Then we must have $a_{j} \leqslant 9$ by Claim 2 and the minimality of $a_{j}$. It follows that $a_{j} \in\{4,7\}$, so $a_{m}=2<a_{j}$ for some $m>j$, contradicting the minimality of $a_{j}$. Therefore, we must have $a_{j} \equiv-1(\bmod 3)$.

It follows from the previous claims that if $a_{0}$ is a multiple of 3 the sequence will eventually reach the periodic pattern $3,6,9,3,6,9, \ldots$; if $a_{0} \equiv-1(\bmod 3)$ the sequence will be strictly increasing; and if $a_{0} \equiv 1(\bmod 3)$ the sequence will be eventually strictly increasing.

So the sequence will be eventually periodic if, and only if, $a_{0}$ is a multiple of 3 .

N2. Let $p \geqslant 2$ be a prime number. Eduardo and Fernando play the following game making moves alternately: in each move, the current player chooses an index $i$ in the set $\{0,1, \ldots, p-1\}$ that was not chosen before by either of the two players and then chooses an element $a_{i}$ of the set $\{0,1,2,3,4,5,6,7,8,9\}$. Eduardo has the first move. The game ends after all the indices $i \in\{0,1, \ldots, p-1\}$ have been chosen. Then the following number is computed:

$$
M=a_{0}+10 \cdot a_{1}+\cdots+10^{p-1} \cdot a_{p-1}=\sum_{j=0}^{p-1} a_{j} \cdot 10^{j}
$$

The goal of Eduardo is to make the number $M$ divisible by $p$, and the goal of Fernando is to prevent this.

Prove that Eduardo has a winning strategy.
(Morocco)
Solution. We say that a player makes the move $\left(i, a_{i}\right)$ if he chooses the index $i$ and then the element $a_{i}$ of the set $\{0,1,2,3,4,5,6,7,8,9\}$ in this move.

If $p=2$ or $p=5$ then Eduardo chooses $i=0$ and $a_{0}=0$ in the first move, and wins, since, independently of the next moves, $M$ will be a multiple of 10 .

Now assume that the prime number $p$ does not belong to $\{2,5\}$. Eduardo chooses $i=p-1$ and $a_{p-1}=0$ in the first move. By Fermat's Little Theorem, $\left(10^{(p-1) / 2}\right)^{2}=10^{p-1} \equiv 1(\bmod p)$, so $p \mid\left(10^{(p-1) / 2}\right)^{2}-1=\left(10^{(p-1) / 2}+1\right)\left(10^{(p-1) / 2}-1\right)$. Since $p$ is prime, either $p \mid 10^{(p-1) / 2}+1$ or $p \mid 10^{(p-1) / 2}-1$. Thus we have two cases:
Case a: $10^{(p-1) / 2} \equiv-1(\bmod p)$
In this case, for each move $\left(i, a_{i}\right)$ of Fernando, Eduardo immediately makes the move $\left(j, a_{j}\right)=$ $\left(i+\frac{p-1}{2}, a_{i}\right)$, if $0 \leqslant i \leqslant \frac{p-3}{2}$, or $\left(j, a_{j}\right)=\left(i-\frac{p-1}{2}, a_{i}\right)$, if $\frac{p-1}{2} \leqslant i \leqslant p-2$. We will have $10^{j} \equiv-10^{i}$ $(\bmod p)$, and so $a_{j} \cdot 10^{j}=a_{i} \cdot 10^{j} \equiv-a_{i} \cdot 10^{i}(\bmod p)$. Notice that this move by Eduardo is always possible. Indeed, immediately before a move by Fernando, for any set of the type $\{r, r+(p-1) / 2\}$ with $0 \leqslant r \leqslant(p-3) / 2$, either no element of this set was chosen as an index by the players in the previous moves or else both elements of this set were chosen as indices by the players in the previous moves. Therefore, after each of his moves, Eduardo always makes the sum of the numbers $a_{k} \cdot 10^{k}$ corresponding to the already chosen pairs ( $k, a_{k}$ ) divisible by $p$, and thus wins the game.
Case b: $10^{(p-1) / 2} \equiv 1(\bmod p)$
In this case, for each move $\left(i, a_{i}\right)$ of Fernando, Eduardo immediately makes the move $\left(j, a_{j}\right)=$ $\left(i+\frac{p-1}{2}, 9-a_{i}\right)$, if $0 \leqslant i \leqslant \frac{p-3}{2}$, or $\left(j, a_{j}\right)=\left(i-\frac{p-1}{2}, 9-a_{i}\right)$, if $\frac{p-1}{2} \leqslant i \leqslant p-2$. The same argument as above shows that Eduardo can always make such move. We will have $10^{j} \equiv 10^{i}$ $(\bmod p)$, and so $a_{j} \cdot 10^{j}+a_{i} \cdot 10^{i} \equiv\left(a_{i}+a_{j}\right) \cdot 10^{i}=9 \cdot 10^{i}(\bmod p)$. Therefore, at the end of the game, the sum of all terms $a_{k} \cdot 10^{k}$ will be congruent to

$$
\sum_{i=0}^{\frac{p-3}{2}} 9 \cdot 10^{i}=10^{(p-1) / 2}-1 \equiv 0 \quad(\bmod p),
$$

and Eduardo wins the game.

N3. Determine all integers $n \geqslant 2$ with the following property: for any integers $a_{1}, a_{2}, \ldots, a_{n}$ whose sum is not divisible by $n$, there exists an index $1 \leqslant i \leqslant n$ such that none of the numbers

$$
a_{i}, a_{i}+a_{i+1}, \ldots, a_{i}+a_{i+1}+\cdots+a_{i+n-1}
$$

is divisible by $n$. (We let $a_{i}=a_{i-n}$ when $i>n$.)
(Thailand)
Answer: These integers are exactly the prime numbers.
Solution. Let us first show that, if $n=a b$, with $a, b \geqslant 2$ integers, then the property in the statement of the problem does not hold. Indeed, in this case, let $a_{k}=a$ for $1 \leqslant k \leqslant n-1$ and $a_{n}=0$. The sum $a_{1}+a_{2}+\cdots+a_{n}=a \cdot(n-1)$ is not divisible by $n$. Let $i$ with $1 \leqslant i \leqslant n$ be an arbitrary index. Taking $j=b$ if $1 \leqslant i \leqslant n-b$, and $j=b+1$ if $n-b<i \leqslant n$, we have

$$
a_{i}+a_{i+1}+\cdots+a_{i+j-1}=a \cdot b=n \equiv 0 \quad(\bmod n) .
$$

It follows that the given example is indeed a counterexample to the property of the statement.
Now let $n$ be a prime number. Suppose by contradiction that the property in the statement of the problem does not hold. Then there are integers $a_{1}, a_{2}, \ldots, a_{n}$ whose sum is not divisible by $n$ such that for each $i, 1 \leqslant i \leqslant n$, there is $j, 1 \leqslant j \leqslant n$, for which the number $a_{i}+a_{i+1}+$ $\cdots+a_{i+j-1}$ is divisible by $n$. Notice that, in any such case, we should have $1 \leqslant j \leqslant n-1$, since $a_{1}+a_{2}+\cdots+a_{n}$ is not divisible by $n$. So we may construct recursively a finite sequence of integers $0=i_{0}<i_{1}<i_{2}<\cdots<i_{n}$ with $i_{s+1}-i_{s} \leqslant n-1$ for $0 \leqslant s \leqslant n-1$ such that, for $0 \leqslant s \leqslant n-1$,

$$
a_{i_{s}+1}+a_{i_{s}+2}+\cdots+a_{i_{s+1}} \equiv 0 \quad(\bmod n)
$$

(where we take indices modulo $n$ ). Indeed, for $0 \leqslant s<n$, we apply the previous observation to $i=i_{s}+1$ in order to define $i_{s+1}=i_{s}+j$.

In the sequence of $n+1$ indices $i_{0}, i_{1}, i_{2}, \ldots, i_{n}$, by the pigeonhole principle, we have two distinct elements which are congruent modulo $n$. So there are indices $r, s$ with $0 \leqslant r<s \leqslant n$ such that $i_{s} \equiv i_{r}(\bmod n)$ and

$$
a_{i_{r}+1}+a_{i_{r}+2}+\cdots+a_{i_{s}}=\sum_{j=r}^{s-1}\left(a_{i_{j}+1}+a_{i_{j}+2}+\cdots+a_{i_{j+1}}\right) \equiv 0 \quad(\bmod n) .
$$

Since $i_{s} \equiv i_{r}(\bmod n)$, we have $i_{s}-i_{r}=k \cdot n$ for some positive integer $k$, and, since $i_{j+1}-i_{j} \leqslant$ $n-1$ for $0 \leqslant j \leqslant n-1$, we have $i_{s}-i_{r} \leqslant(n-1) \cdot n$, so $k \leqslant n-1$. But in this case

$$
a_{i_{r}+1}+a_{i_{r}+2}+\cdots+a_{i_{s}}=k \cdot\left(a_{1}+a_{2}+\cdots+a_{n}\right)
$$

cannot be a multiple of $n$, since $n$ is prime and neither $k$ nor $a_{1}+a_{2}+\cdots+a_{n}$ is a multiple of $n$. A contradiction.

N4. Call a rational number short if it has finitely many digits in its decimal expansion. For a positive integer $m$, we say that a positive integer $t$ is $m$-tastic if there exists a number $c \in\{1,2,3, \ldots, 2017\}$ such that $\frac{10^{t}-1}{c \cdot m}$ is short, and such that $\frac{10^{k}-1}{c \cdot m}$ is not short for any $1 \leqslant k<t$. Let $S(m)$ be the set of $m$-tastic numbers. Consider $S(m)$ for $m=1,2, \ldots$. What is the maximum number of elements in $S(m)$ ?
(Turkey)
Answer: 807.
Solution. First notice that $x \in \mathbb{Q}$ is short if and only if there are exponents $a, b \geqslant 0$ such that $2^{a} \cdot 5^{b} \cdot x \in \mathbb{Z}$. In fact, if $x$ is short, then $x=\frac{n}{10^{k}}$ for some $k$ and we can take $a=b=k$; on the other hand, if $2^{a} \cdot 5^{b} \cdot x=q \in \mathbb{Z}$ then $x=\frac{2^{b} \cdot 5^{a} q}{10^{a+b}}$, so $x$ is short.

If $m=2^{a} \cdot 5^{b} \cdot s$, with $\operatorname{gcd}(s, 10)=1$, then $\frac{10^{t}-1}{m}$ is short if and only if $s$ divides $10^{t}-1$. So we may (and will) suppose without loss of generality that $\operatorname{gcd}(m, 10)=1$. Define

$$
C=\{1 \leqslant c \leqslant 2017: \operatorname{gcd}(c, 10)=1\} .
$$

The $m$-tastic numbers are then precisely the smallest exponents $t>0$ such that $10^{t} \equiv 1$ $(\bmod c m)$ for some integer $c \in C$, that is, the set of orders of 10 modulo cm . In other words,

$$
S(m)=\left\{\operatorname{ord}_{c m}(10): c \in C\right\} .
$$

Since there are $4 \cdot 201+3=807$ numbers $c$ with $1 \leqslant c \leqslant 2017$ and $\operatorname{gcd}(c, 10)=1$, namely those such that $c \equiv 1,3,7,9(\bmod 10)$,

$$
|S(m)| \leqslant|C|=807
$$

Now we find $m$ such that $|S(m)|=807$. Let

$$
P=\{1<p \leqslant 2017: p \text { is prime, } p \neq 2,5\}
$$

and choose a positive integer $\alpha$ such that every $p \in P$ divides $10^{\alpha}-1$ (e.g. $\alpha=\varphi(T), T$ being the product of all primes in $P$ ), and let $m=10^{\alpha}-1$.
Claim. For every $c \in C$, we have

$$
\operatorname{ord}_{c m}(10)=c \alpha
$$

As an immediate consequence, this implies $|S(m)|=|C|=807$, finishing the problem.
Proof. Obviously $\operatorname{ord}_{m}(10)=\alpha$. Let $t=\operatorname{ord}_{c m}(10)$. Then

$$
c m\left|10^{t}-1 \quad \Longrightarrow \quad m\right| 10^{t}-1 \quad \Longrightarrow \quad \alpha \mid t
$$

Hence $t=k \alpha$ for some $k \in \mathbb{Z}_{>0}$. We will show that $k=c$.
Denote by $\nu_{p}(n)$ the number of prime factors $p$ in $n$, that is, the maximum exponent $\beta$ for which $p^{\beta} \mid n$. For every $\ell \geqslant 1$ and $p \in P$, the Lifting the Exponent Lemma provides

$$
\nu_{p}\left(10^{\ell \alpha}-1\right)=\nu_{p}\left(\left(10^{\alpha}\right)^{\ell}-1\right)=\nu_{p}\left(10^{\alpha}-1\right)+\nu_{p}(\ell)=\nu_{p}(m)+\nu_{p}(\ell)
$$

so

$$
\begin{aligned}
c m \mid 10^{k \alpha}-1 & \Longleftrightarrow \forall p \in P ; \nu_{p}(c m) \leqslant \nu_{p}\left(10^{k \alpha}-1\right) \\
& \Longleftrightarrow \forall p \in P ; \nu_{p}(m)+\nu_{p}(c) \leqslant \nu_{p}(m)+\nu_{p}(k) \\
& \Longleftrightarrow \forall p \in P ; \nu_{p}(c) \leqslant \nu_{p}(k) \\
& \Longleftrightarrow c \mid k .
\end{aligned}
$$

The first such $k$ is $k=c$, so $\operatorname{ord}_{c m}(10)=c \alpha$.

Comment. The Lifting the Exponent Lemma states that, for any odd prime $p$, any integers $a, b$ coprime with $p$ such that $p \mid a-b$, and any positive integer exponent $n$,

$$
\nu_{p}\left(a^{n}-b^{n}\right)=\nu_{p}(a-b)+\nu_{p}(n),
$$

and, for $p=2$,

$$
\nu_{2}\left(a^{n}-b^{n}\right)=\nu_{2}\left(a^{2}-b^{2}\right)+\nu_{p}(n)-1 .
$$

Both claims can be proved by induction on $n$.

N5. Find all pairs $(p, q)$ of prime numbers with $p>q$ for which the number

$$
\frac{(p+q)^{p+q}(p-q)^{p-q}-1}{(p+q)^{p-q}(p-q)^{p+q}-1}
$$

is an integer.
(Japan)
Answer: The only such pair is $(3,2)$.
Solution. Let $M=(p+q)^{p-q}(p-q)^{p+q}-1$, which is relatively prime with both $p+q$ and $p-q$. Denote by $(p-q)^{-1}$ the multiplicative inverse of $(p-q)$ modulo $M$.

By eliminating the term -1 in the numerator,

$$
\begin{align*}
(p+q)^{p+q}(p-q)^{p-q}-1 & \equiv(p+q)^{p-q}(p-q)^{p+q}-1 \quad(\bmod M) \\
(p+q)^{2 q} & \equiv(p-q)^{2 q} \quad(\bmod M)  \tag{1}\\
\left((p+q) \cdot(p-q)^{-1}\right)^{2 q} & \equiv 1 \quad(\bmod M) \tag{2}
\end{align*}
$$

Case 1: $q \geqslant 5$.
Consider an arbitrary prime divisor $r$ of $M$. Notice that $M$ is odd, so $r \geqslant 3$. By (2), the multiplicative order of $\left((p+q) \cdot(p-q)^{-1}\right)$ modulo $r$ is a divisor of the exponent $2 q$ in (2), so it can be $1,2, q$ or $2 q$.

By Fermat's theorem, the order divides $r-1$. So, if the order is $q$ or $2 q$ then $r \equiv 1(\bmod q)$. If the order is 1 or 2 then $r \mid(p+q)^{2}-(p-q)^{2}=4 p q$, so $r=p$ or $r=q$. The case $r=p$ is not possible, because, by applying Fermat's theorem,
$M=(p+q)^{p-q}(p-q)^{p+q}-1 \equiv q^{p-q}(-q)^{p+q}-1=\left(q^{2}\right)^{p}-1 \equiv q^{2}-1=(q+1)(q-1) \quad(\bmod p)$
and the last factors $q-1$ and $q+1$ are less than $p$ and thus $p \nmid M$. Hence, all prime divisors of $M$ are either $q$ or of the form $k q+1$; it follows that all positive divisors of $M$ are congruent to 0 or 1 modulo $q$.

Now notice that

$$
M=\left((p+q)^{\frac{p-q}{2}}(p-q)^{\frac{p+q}{2}}-1\right)\left((p+q)^{\frac{p-q}{2}}(p-q)^{\frac{p+q}{2}}+1\right)
$$

is the product of two consecutive positive odd numbers; both should be congruent to 0 or 1 modulo $q$. But this is impossible by the assumption $q \geqslant 5$. So, there is no solution in Case 1 .
Case 2: $q=2$.
By (1), we have $M \mid(p+q)^{2 q}-(p-q)^{2 q}=(p+2)^{4}-(p-2)^{4}$, so

$$
\begin{gathered}
(p+2)^{p-2}(p-2)^{p+2}-1=M \leqslant(p+2)^{4}-(p-2)^{4} \leqslant(p+2)^{4}-1, \\
(p+2)^{p-6}(p-2)^{p+2} \leqslant 1 .
\end{gathered}
$$

If $p \geqslant 7$ then the left-hand side is obviously greater than 1 . For $p=5$ we have $(p+2)^{p-6}(p-2)^{p+2}=7^{-1} \cdot 3^{7}$ which is also too large.

There remains only one candidate, $p=3$, which provides a solution:

$$
\frac{(p+q)^{p+q}(p-q)^{p-q}-1}{(p+q)^{p-q}(p-q)^{p+q}-1}=\frac{5^{5} \cdot 1^{1}-1}{5^{1} \cdot 1^{5}-1}=\frac{3124}{4}=781 .
$$

So in Case 2 the only solution is $(p, q)=(3,2)$.

Case 3: $q=3$.
Similarly to Case 2, we have

$$
M \left\lvert\,(p+q)^{2 q}-(p-q)^{2 q}=64 \cdot\left(\left(\frac{p+3}{2}\right)^{6}-\left(\frac{p-3}{2}\right)^{6}\right) .\right.
$$

Since $M$ is odd, we conclude that

$$
M \left\lvert\,\left(\frac{p+3}{2}\right)^{6}-\left(\frac{p-3}{2}\right)^{6}\right.
$$

and

$$
\begin{gathered}
(p+3)^{p-3}(p-3)^{p+3}-1=M \leqslant\left(\frac{p+3}{2}\right)^{6}-\left(\frac{p-3}{2}\right)^{6} \leqslant\left(\frac{p+3}{2}\right)^{6}-1, \\
64(p+3)^{p-9}(p-3)^{p+3} \leqslant 1
\end{gathered}
$$

If $p \geqslant 11$ then the left-hand side is obviously greater than 1 . If $p=7$ then the left-hand side is $64 \cdot 10^{-2} \cdot 4^{10}>1$. If $p=5$ then the left-hand side is $64 \cdot 8^{-4} \cdot 2^{8}=2^{2}>1$. Therefore, there is no solution in Case 3.

N6. Find the smallest positive integer $n$, or show that no such $n$ exists, with the following property: there are infinitely many distinct $n$-tuples of positive rational numbers ( $a_{1}, a_{2}, \ldots, a_{n}$ ) such that both

$$
a_{1}+a_{2}+\cdots+a_{n} \quad \text { and } \quad \frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}
$$

are integers.
(Singapore)
Answer: $n=3$.
Solution 1. For $n=1, a_{1} \in \mathbb{Z}_{>0}$ and $\frac{1}{a_{1}} \in \mathbb{Z}_{>0}$ if and only if $a_{1}=1$. Next we show that
(i) There are finitely many $(x, y) \in \mathbb{Q}_{>0}^{2}$ satisfying $x+y \in \mathbb{Z}$ and $\frac{1}{x}+\frac{1}{y} \in \mathbb{Z}$

Write $x=\frac{a}{b}$ and $y=\frac{c}{d}$ with $a, b, c, d \in \mathbb{Z}_{>0}$ and $\operatorname{gcd}(a, b)=\operatorname{gcd}(c, d)=1$. Then $x+y \in \mathbb{Z}$ and $\frac{1}{x}+\frac{1}{y} \in \mathbb{Z}$ is equivalent to the two divisibility conditions

$$
\begin{equation*}
b d \mid a d+b c \quad(1) \quad \text { and } \quad a c \mid a d+b c \tag{2}
\end{equation*}
$$

Condition (1) implies that $d|a d+b c \Longleftrightarrow d| b c \Longleftrightarrow d \mid b$ since $\operatorname{gcd}(c, d)=1$. Still from (1) we get $b|a d+b c \Longleftrightarrow b| a d \Longleftrightarrow b \mid d$ since $\operatorname{gcd}(a, b)=1$. From $b \mid d$ and $d \mid b$ we have $b=d$.
An analogous reasoning with condition (2) shows that $a=c$. Hence $x=\frac{a}{b}=\frac{c}{d}=y$, i.e., the problem amounts to finding all $x \in \mathbb{Q}_{>0}$ such that $2 x \in \mathbb{Z}_{>0}$ and $\frac{2}{x} \in \mathbb{Z}_{>0}$. Letting $n=2 x \in \mathbb{Z}_{>0}$, we have that $\frac{2}{x} \in \mathbb{Z}_{>0} \Longleftrightarrow \frac{4}{n} \in \mathbb{Z}_{>0} \Longleftrightarrow n=1,2$ or 4 , and there are finitely many solutions, namely $(x, y)=\left(\frac{1}{2}, \frac{1}{2}\right),(1,1)$ or $(2,2)$.
(ii) There are infinitely many triples $(x, y, z) \in \mathbb{Q}_{>0}^{2}$ such that $x+y+z \in \mathbb{Z}$ and $\frac{1}{x}+\frac{1}{y}+\frac{1}{z} \in \mathbb{Z}$. We will look for triples such that $x+y+z=1$, so we may write them in the form

$$
(x, y, z)=\left(\frac{a}{a+b+c}, \frac{b}{a+b+c}, \frac{c}{a+b+c}\right) \quad \text { with } a, b, c \in \mathbb{Z}_{>0}
$$

We want these to satisfy

$$
\frac{1}{x}+\frac{1}{y}+\frac{1}{z}=\frac{a+b+c}{a}+\frac{a+b+c}{b}+\frac{a+b+c}{c} \in \mathbb{Z} \Longleftrightarrow \frac{b+c}{a}+\frac{a+c}{b}+\frac{a+b}{c} \in \mathbb{Z}
$$

Fixing $a=1$, it suffices to find infinitely many pairs $(b, c) \in \mathbb{Z}_{>0}^{2}$ such that

$$
\begin{equation*}
\frac{1}{b}+\frac{1}{c}+\frac{c}{b}+\frac{b}{c}=3 \Longleftrightarrow b^{2}+c^{2}-3 b c+b+c=0 \tag{*}
\end{equation*}
$$

To show that equation (*) has infinitely many solutions, we use Vieta jumping (also known as root flipping): starting with $b=2, c=3$, the following algorithm generates infinitely many solutions. Let $c \geqslant b$, and view (*) as a quadratic equation in $b$ for $c$ fixed:

$$
\begin{equation*}
b^{2}-(3 c-1) \cdot b+\left(c^{2}+c\right)=0 \tag{**}
\end{equation*}
$$

Then there exists another root $b_{0} \in \mathbb{Z}$ of $(* *)$ which satisfies $b+b_{0}=3 c-1$ and $b \cdot b_{0}=c^{2}+c$. Since $c \geqslant b$ by assumption,

$$
b_{0}=\frac{c^{2}+c}{b} \geqslant \frac{c^{2}+c}{c}>c
$$

Hence from the solution $(b, c)$ we obtain another one $\left(c, b_{0}\right)$ with $b_{0}>c$, and we can then "jump" again, this time with $c$ as the "variable" in the quadratic (*). This algorithm will generate an infinite sequence of distinct solutions, whose first terms are
$(2,3),(3,6),(6,14),(14,35),(35,90),(90,234),(234,611),(611,1598),(1598,4182), \ldots$

Comment. Although not needed for solving this problem, we may also explicitly solve the recursion given by the Vieta jumping. Define the sequence ( $x_{n}$ ) as follows:

$$
x_{0}=2, \quad x_{1}=3 \quad \text { and } \quad x_{n+2}=3 x_{n+1}-x_{n}-1 \text { for } n \geqslant 0
$$

Then the triple

$$
(x, y, z)=\left(\frac{1}{1+x_{n}+x_{n+1}}, \frac{x_{n}}{1+x_{n}+x_{n+1}}, \frac{x_{n+1}}{1+x_{n}+x_{n+1}}\right)
$$

satisfies the problem conditions for all $n \in \mathbb{N}$. It is easy to show that $x_{n}=F_{2 n+1}+1$, where $F_{n}$ denotes the $n$-th term of the Fibonacci sequence ( $F_{0}=0, F_{1}=1$, and $F_{n+2}=F_{n+1}+F_{n}$ for $n \geqslant 0$ ).
Solution 2. Call the $n$-tuples $\left(a_{1}, a_{2}, \ldots, a_{n}\right) \in \mathbb{Q}_{>0}^{n}$ satisfying the conditions of the problem statement good, and those for which

$$
f\left(a_{1}, \ldots, a_{n}\right) \stackrel{\text { def }}{=}\left(a_{1}+a_{2}+\cdots+a_{n}\right)\left(\frac{1}{a_{1}}+\frac{1}{a_{2}}+\cdots+\frac{1}{a_{n}}\right)
$$

is an integer pretty. Then good $n$-tuples are pretty, and if $\left(b_{1}, \ldots, b_{n}\right)$ is pretty then

$$
\left(\frac{b_{1}}{b_{1}+b_{2}+\cdots+b_{n}}, \frac{b_{2}}{b_{1}+b_{2}+\cdots+b_{n}}, \ldots, \frac{b_{n}}{b_{1}+b_{2}+\cdots+b_{n}}\right)
$$

is good since the sum of its components is 1 , and the sum of the reciprocals of its components equals $f\left(b_{1}, \ldots, b_{n}\right)$. We declare pretty $n$-tuples proportional to each other equivalent since they are precisely those which give rise to the same good $n$-tuple. Clearly, each such equivalence class contains exactly one $n$-tuple of positive integers having no common prime divisors. Call such $n$-tuple a primitive pretty tuple. Our task is to find infinitely many primitive pretty $n$-tuples.

For $n=1$, there is clearly a single primitive 1 -tuple. For $n=2$, we have $f(a, b)=\frac{(a+b)^{2}}{a b}$, which can be integral (for coprime $a, b \in \mathbb{Z}_{>0}$ ) only if $a=b=1$ (see for instance (i) in the first solution).

Now we construct infinitely many primitive pretty triples for $n=3$. Fix $b, c, k \in \mathbb{Z}_{>0}$; we will try to find sufficient conditions for the existence of an $a \in \mathbb{Q}_{>0}$ such that $f(a, b, c)=k$. Write $\sigma=b+c, \tau=b c$. From $f(a, b, c)=k$, we have that $a$ should satisfy the quadratic equation

$$
\begin{equation*}
a^{2} \cdot \sigma+a \cdot\left(\sigma^{2}-(k-1) \tau\right)+\sigma \tau=0 \tag{1}
\end{equation*}
$$

whose discriminant is

$$
\Delta=\left(\sigma^{2}-(k-1) \tau\right)^{2}-4 \sigma^{2} \tau=\left((k+1) \tau-\sigma^{2}\right)^{2}-4 k \tau^{2}
$$

We need it to be a square of an integer, say, $\Delta=M^{2}$ for some $M \in \mathbb{Z}$, i.e., we want

$$
\left((k+1) \tau-\sigma^{2}\right)^{2}-M^{2}=2 k \cdot 2 \tau^{2}
$$

so that it suffices to set

$$
(k+1) \tau-\sigma^{2}=\tau^{2}+k, \quad M=\tau^{2}-k .
$$

The first relation reads $\sigma^{2}=(\tau-1)(k-\tau)$, so if $b$ and $c$ satisfy

$$
\begin{equation*}
\tau-1 \mid \sigma^{2} \quad \text { i.e. } \quad b c-1 \mid(b+c)^{2} \tag{2}
\end{equation*}
$$

then $k=\frac{\sigma^{2}}{\tau-1}+\tau$ will be integral, and we find rational solutions to (1), namely

$$
a=\frac{\sigma}{\tau-1}=\frac{b+c}{b c-1} \quad \text { or } \quad a=\frac{\tau^{2}-\tau}{\sigma}=\frac{b c \cdot(b c-1)}{b+c}
$$

We can now find infinitely many pairs ( $b, c$ ) satisfying (2) by Vieta jumping. For example, if we impose

$$
(b+c)^{2}=5 \cdot(b c-1)
$$

then all pairs $(b, c)=\left(v_{i}, v_{i+1}\right)$ satisfy the above condition, where

$$
v_{1}=2, v_{2}=3, \quad v_{i+2}=3 v_{i+1}-v_{i} \quad \text { for } i \geqslant 0
$$

For $(b, c)=\left(v_{i}, v_{i+1}\right)$, one of the solutions to (1) will be $a=(b+c) /(b c-1)=5 /(b+c)=$ $5 /\left(v_{i}+v_{i+1}\right)$. Then the pretty triple ( $a, b, c$ ) will be equivalent to the integral pretty triple

$$
\left(5, v_{i}\left(v_{i}+v_{i+1}\right), v_{i+1}\left(v_{i}+v_{i+1}\right)\right)
$$

After possibly dividing by 5 , we obtain infinitely many primitive pretty triples, as required.
Comment. There are many other infinite series of $(b, c)=\left(v_{i}, v_{i+1}\right)$ with $b c-1 \mid(b+c)^{2}$. Some of them are:

$$
\begin{array}{llll}
v_{1}=1, & v_{2}=3, & v_{i+1}=6 v_{i}-v_{i-1}, & \left(v_{i}+v_{i+1}\right)^{2}=8 \cdot\left(v_{i} v_{i+1}-1\right) ; \\
v_{1}=1, & v_{2}=2, & v_{i+1}=7 v_{i}-v_{i-1}, & \left(v_{i}+v_{i+1}\right)^{2}=9 \cdot\left(v_{i} v_{i+1}-1\right) ; \\
v_{1}=1, & v_{2}=5, & v_{i+1}=7 v_{i}-v_{i-1}, & \left(v_{i}+v_{i+1}\right)^{2}=9 \cdot\left(v_{i} v_{i+1}-1\right)
\end{array}
$$

(the last two are in fact one sequence prolonged in two possible directions).

N7. Say that an ordered pair $(x, y)$ of integers is an irreducible lattice point if $x$ and $y$ are relatively prime. For any finite set $S$ of irreducible lattice points, show that there is a homogenous polynomial in two variables, $f(x, y)$, with integer coefficients, of degree at least 1 , such that $f(x, y)=1$ for each $(x, y)$ in the set $S$.

Note: A homogenous polynomial of degree $n$ is any nonzero polynomial of the form

$$
\begin{equation*}
f(x, y)=a_{0} x^{n}+a_{1} x^{n-1} y+a_{2} x^{n-2} y^{2}+\cdots+a_{n-1} x y^{n-1}+a_{n} y^{n} . \tag{U.S.A.}
\end{equation*}
$$

Solution 1. First of all, we note that finding a homogenous polynomial $f(x, y)$ such that $f(x, y)= \pm 1$ is enough, because we then have $f^{2}(x, y)=1$. Label the irreducible lattice points $\left(x_{1}, y_{1}\right)$ through $\left(x_{n}, y_{n}\right)$. If any two of these lattice points $\left(x_{i}, y_{i}\right)$ and $\left(x_{j}, y_{j}\right)$ lie on the same line through the origin, then $\left(x_{j}, y_{j}\right)=\left(-x_{i},-y_{i}\right)$ because both of the points are irreducible. We then have $f\left(x_{j}, y_{j}\right)= \pm f\left(x_{i}, y_{i}\right)$ whenever $f$ is homogenous, so we can assume that no two of the lattice points are collinear with the origin by ignoring the extra lattice points.

Consider the homogenous polynomials $\ell_{i}(x, y)=y_{i} x-x_{i} y$ and define

$$
g_{i}(x, y)=\prod_{j \neq i} \ell_{j}(x, y)
$$

Then $\ell_{i}\left(x_{j}, y_{j}\right)=0$ if and only if $j=i$, because there is only one lattice point on each line through the origin. Thus, $g_{i}\left(x_{j}, y_{j}\right)=0$ for all $j \neq i$. Define $a_{i}=g_{i}\left(x_{i}, y_{i}\right)$, and note that $a_{i} \neq 0$.

Note that $g_{i}(x, y)$ is a degree $n-1$ polynomial with the following two properties:

1. $g_{i}\left(x_{j}, y_{j}\right)=0$ if $j \neq i$.
2. $g_{i}\left(x_{i}, y_{i}\right)=a_{i}$.

For any $N \geqslant n-1$, there also exists a polynomial of degree $N$ with the same two properties. Specifically, let $I_{i}(x, y)$ be a degree 1 homogenous polynomial such that $I_{i}\left(x_{i}, y_{i}\right)=1$, which exists since $\left(x_{i}, y_{i}\right)$ is irreducible. Then $I_{i}(x, y)^{N-(n-1)} g_{i}(x, y)$ satisfies both of the above properties and has degree $N$.

We may now reduce the problem to the following claim:
Claim: For each positive integer a, there is a homogenous polynomial $f_{a}(x, y)$, with integer coefficients, of degree at least 1 , such that $f_{a}(x, y) \equiv 1(\bmod a)$ for all relatively prime $(x, y)$.

To see that this claim solves the problem, take $a$ to be the least common multiple of the numbers $a_{i}(1 \leqslant i \leqslant n)$. Take $f_{a}$ given by the claim, choose some power $f_{a}(x, y)^{k}$ that has degree at least $n-1$, and subtract appropriate multiples of the $g_{i}$ constructed above to obtain the desired polynomial.

We prove the claim by factoring $a$. First, if $a$ is a power of a prime ( $a=p^{k}$ ), then we may choose either:

- $f_{a}(x, y)=\left(x^{p-1}+y^{p-1}\right)^{\phi(a)}$ if $p$ is odd;
- $f_{a}(x, y)=\left(x^{2}+x y+y^{2}\right)^{\phi(a)}$ if $p=2$.

Now suppose $a$ is any positive integer, and let $a=q_{1} q_{2} \cdots q_{k}$, where the $q_{i}$ are prime powers, pairwise relatively prime. Let $f_{q_{i}}$ be the polynomials just constructed, and let $F_{q_{i}}$ be powers of these that all have the same degree. Note that

$$
\frac{a}{q_{i}} F_{q_{i}}(x, y) \equiv \frac{a}{q_{i}} \quad(\bmod a)
$$

for any relatively prime $x, y$. By Bézout's lemma, there is an integer linear combination of the $\frac{a}{q_{i}}$ that equals 1. Thus, there is a linear combination of the $F_{q_{i}}$ such that $F_{q_{i}}(x, y) \equiv 1$ $(\bmod a)$ for any relatively prime $(x, y)$; and this polynomial is homogenous because all the $F_{q_{i}}$ have the same degree.

Solution 2. As in the previous solution, label the irreducible lattice points $\left(x_{1}, y_{1}\right), \ldots,\left(x_{n}, y_{n}\right)$ and assume without loss of generality that no two of the points are collinear with the origin. We induct on $n$ to construct a homogenous polynomial $f(x, y)$ such that $f\left(x_{i}, y_{i}\right)=1$ for all $1 \leqslant i \leqslant n$.

If $n=1$ : Since $x_{1}$ and $y_{1}$ are relatively prime, there exist some integers $c, d$ such that $c x_{1}+d y_{1}=1$. Then $f(x, y)=c x+d y$ is suitable.

If $n \geqslant 2$ : By the induction hypothesis we already have a homogeneous polynomial $g(x, y)$ with $g\left(x_{1}, y_{1}\right)=\ldots=g\left(x_{n-1}, y_{n-1}\right)=1$. Let $j=\operatorname{deg} g$,

$$
g_{n}(x, y)=\prod_{k=1}^{n-1}\left(y_{k} x-x_{k} y\right)
$$

and $a_{n}=g_{n}\left(x_{n}, y_{n}\right)$. By assumption, $a_{n} \neq 0$. Take some integers $c, d$ such that $c x_{n}+d y_{n}=1$. We will construct $f(x, y)$ in the form

$$
f(x, y)=g(x, y)^{K}-C \cdot g_{n}(x, y) \cdot(c x+d y)^{L}
$$

where $K$ and $L$ are some positive integers and $C$ is some integer. We assume that $L=K j-n+1$ so that $f$ is homogenous.

Due to $g\left(x_{1}, y_{1}\right)=\ldots=g\left(x_{n-1}, y_{n-1}\right)=1$ and $g_{n}\left(x_{1}, y_{1}\right)=\ldots=g_{n}\left(x_{n-1}, y_{n-1}\right)=0$, the property $f\left(x_{1}, y_{1}\right)=\ldots=f\left(x_{n-1}, y_{n-1}\right)=1$ is automatically satisfied with any choice of $K, L$, and $C$.

Furthermore,

$$
f\left(x_{n}, y_{n}\right)=g\left(x_{n}, y_{n}\right)^{K}-C \cdot g_{n}\left(x_{n}, y_{n}\right) \cdot\left(c x_{n}+d y_{n}\right)^{L}=g\left(x_{n}, y_{n}\right)^{K}-C a_{n} .
$$

If we have an exponent $K$ such that $g\left(x_{n}, y_{n}\right)^{K} \equiv 1\left(\bmod a_{n}\right)$, then we may choose $C$ such that $f\left(x_{n}, y_{n}\right)=1$. We now choose such a $K$.

Consider an arbitrary prime divisor $p$ of $a_{n}$. By

$$
p \mid a_{n}=g_{n}\left(x_{n}, y_{n}\right)=\prod_{k=1}^{n-1}\left(y_{k} x_{n}-x_{k} y_{n}\right)
$$

there is some $1 \leqslant k<n$ such that $x_{k} y_{n} \equiv x_{n} y_{k}(\bmod p)$. We first show that $x_{k} x_{n}$ or $y_{k} y_{n}$ is relatively prime with $p$. This is trivial in the case $x_{k} y_{n} \equiv x_{n} y_{k} \not \equiv 0(\bmod p)$. In the other case, we have $x_{k} y_{n} \equiv x_{n} y_{k} \equiv 0(\bmod p)$, If, say $p \mid x_{k}$, then $p \nmid y_{k}$ because $\left(x_{k}, y_{k}\right)$ is irreducible, so $p \mid x_{n}$; then $p \nmid y_{n}$ because $\left(x_{k}, y_{k}\right)$ is irreducible. In summary, $p \mid x_{k}$ implies $p \nmid y_{k} y_{n}$. Similarly, $p \mid y_{n}$ implies $p \nmid x_{k} x_{n}$.

By the homogeneity of $g$ we have the congruences

$$
\begin{equation*}
x_{k}^{d} \cdot g\left(x_{n}, y_{n}\right)=g\left(x_{k} x_{n}, x_{k} y_{n}\right) \equiv g\left(x_{k} x_{n}, y_{k} x_{n}\right)=x_{n}^{d} \cdot g\left(x_{k}, y_{k}\right)=x_{n}^{d} \quad(\bmod p) \tag{1.1}
\end{equation*}
$$

and

$$
\begin{equation*}
y_{k}^{d} \cdot g\left(x_{n}, y_{n}\right)=g\left(y_{k} x_{n}, y_{k} y_{n}\right) \equiv g\left(x_{k} y_{n}, y_{k} y_{n}\right)=y_{n}^{d} \cdot g\left(x_{k}, y_{k}\right)=y_{n}^{d} \quad(\bmod p) \tag{1.2}
\end{equation*}
$$

If $p \nmid x_{k} x_{n}$, then take the $(p-1)^{s t}$ power of (1.1); otherwise take the $(p-1)^{s t}$ power of (1.2); by Fermat's theorem, in both cases we get

$$
g\left(x_{n}, y_{n}\right)^{p-1} \equiv 1 \quad(\bmod p) .
$$

If $p^{\alpha} \mid m$, then we have

$$
g\left(x_{n}, y_{n}\right)^{p^{\alpha-1}(p-1)} \equiv 1 \quad\left(\bmod p^{\alpha}\right)
$$

which implies that the exponent $K=n \cdot \varphi\left(a_{n}\right)$, which is a multiple of all $p^{\alpha-1}(p-1)$, is a suitable choice. (The factor $n$ is added only so that $K \geqslant n$ and so $L>0$.)

Comment. It is possible to show that there is no constant $C$ for which, given any two irreducible lattice points, there is some homogenous polynomial $f$ of degree at most $C$ with integer coefficients that takes the value 1 on the two points. Indeed, if one of the points is $(1,0)$ and the other is $(a, b)$, the polynomial $f(x, y)=a_{0} x^{n}+a_{1} x^{n-1} y+\cdots+a_{n} y^{n}$ should satisfy $a_{0}=1$, and so $a^{n} \equiv 1(\bmod b)$. If $a=3$ and $b=2^{k}$ with $k \geqslant 3$, then $n \geqslant 2^{k-2}$. If we choose $2^{k-2}>C$, this gives a contradiction.

N8. Let $p$ be an odd prime number and $\mathbb{Z}_{>0}$ be the set of positive integers. Suppose that a function $f: \mathbb{Z}_{>0} \times \mathbb{Z}_{>0} \rightarrow\{0,1\}$ satisfies the following properties:

- $f(1,1)=0$;
- $f(a, b)+f(b, a)=1$ for any pair of relatively prime positive integers $(a, b)$ not both equal to 1 ;
- $f(a+b, b)=f(a, b)$ for any pair of relatively prime positive integers $(a, b)$.

Prove that

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right) \geqslant \sqrt{2 p}-2
$$

(Italy)
Solution 1. Denote by $\mathbb{A}$ the set of all pairs of coprime positive integers. Notice that for every $(a, b) \in \mathbb{A}$ there exists a pair $(u, v) \in \mathbb{Z}^{2}$ with $u a+v b=1$. Moreover, if $\left(u_{0}, v_{0}\right)$ is one such pair, then all such pairs are of the form $(u, v)=\left(u_{0}+k b, v_{0}-k a\right)$, where $k \in \mathbb{Z}$. So there exists a unique such pair $(u, v)$ with $-b / 2<u \leqslant b / 2$; we denote this pair by $(u, v)=g(a, b)$.
Lemma. Let $(a, b) \in \mathbb{A}$ and $(u, v)=g(a, b)$. Then $f(a, b)=1 \Longleftrightarrow u>0$.
Proof. We induct on $a+b$. The base case is $a+b=2$. In this case, we have that $a=b=1$, $g(a, b)=g(1,1)=(0,1)$ and $f(1,1)=0$, so the claim holds.

Assume now that $a+b>2$, and so $a \neq b$, since $a$ and $b$ are coprime. Two cases are possible. Case 1: $a>b$.

Notice that $g(a-b, b)=(u, v+u)$, since $u(a-b)+(v+u) b=1$ and $u \in(-b / 2, b / 2]$. Thus $f(a, b)=1 \Longleftrightarrow f(a-b, b)=1 \Longleftrightarrow u>0$ by the induction hypothesis.
Case 2: $a<b$. (Then, clearly, $b \geqslant 2$.)
Now we estimate $v$. Since $v b=1-u a$, we have

$$
1+\frac{a b}{2}>v b \geqslant 1-\frac{a b}{2}, \quad \text { so } \quad \frac{1+a}{2} \geqslant \frac{1}{b}+\frac{a}{2}>v \geqslant \frac{1}{b}-\frac{a}{2}>-\frac{a}{2} .
$$

Thus $1+a>2 v>-a$, so $a \geqslant 2 v>-a$, hence $a / 2 \geqslant v>-a / 2$, and thus $g(b, a)=(v, u)$.
Observe that $f(a, b)=1 \Longleftrightarrow f(b, a)=0 \Longleftrightarrow f(b-a, a)=0$. We know from Case 1 that $g(b-a, a)=(v, u+v)$. We have $f(b-a, a)=0 \Longleftrightarrow v \leqslant 0$ by the inductive hypothesis. Then, since $b>a \geqslant 1$ and $u a+v b=1$, we have $v \leqslant 0 \Longleftrightarrow u>0$, and we are done.

The Lemma proves that, for all $(a, b) \in \mathbb{A}, f(a, b)=1$ if and only if the inverse of $a$ modulo $b$, taken in $\{1,2, \ldots, b-1\}$, is at most $b / 2$. Then, for any odd prime $p$ and integer $n$ such that $n \not \equiv 0(\bmod p), f\left(n^{2}, p\right)=1$ iff the inverse of $n^{2} \bmod p$ is less than $p / 2$. Since $\left\{n^{2} \bmod p: 1 \leqslant n \leqslant p-1\right\}=\left\{n^{-2} \bmod p: 1 \leqslant n \leqslant p-1\right\}$, including multiplicities (two for each quadratic residue in each set), we conclude that the desired sum is twice the number of quadratic residues that are less than $p / 2$, i.e.,

$$
\begin{equation*}
\left.\sum_{n=1}^{p-1} f\left(n^{2}, p\right)=2 \left\lvert\,\left\{k: 1 \leqslant k \leqslant \frac{p-1}{2} \text { and } k^{2} \bmod p<\frac{p}{2}\right\}\right. \right\rvert\, . \tag{1}
\end{equation*}
$$

Since the number of perfect squares in the interval $[1, p / 2)$ is $\lfloor\sqrt{p / 2}\rfloor>\sqrt{p / 2}-1$, we conclude that

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right)>2\left(\sqrt{\frac{p}{2}}-1\right)=\sqrt{2 p}-2 .
$$

Solution 2. We provide a different proof for the Lemma. For this purpose, we use continued fractions to find $g(a, b)=(u, v)$ explicitly.

The function $f$ is completely determined on $\mathbb{A}$ by the following
Claim. Represent $a / b$ as a continued fraction; that is, let $a_{0}$ be an integer and $a_{1}, \ldots, a_{k}$ be positive integers such that $a_{k} \geqslant 2$ and

$$
\frac{a}{b}=a_{0}+\frac{1}{a_{1}+\frac{1}{a_{2}+\frac{1}{\cdots+\frac{1}{a_{k}}}}}=\left[a_{0} ; a_{1}, a_{2}, \ldots, a_{k}\right] .
$$

Then $f(a, b)=0 \Longleftrightarrow k$ is even.
Proof. We induct on $b$. If $b=1$, then $a / b=[a]$ and $k=0$. Then, for $a \geqslant 1$, an easy induction shows that $f(a, 1)=f(1,1)=0$.

Now consider the case $b>1$. Perform the Euclidean division $a=q b+r$, with $0 \leqslant r<b$. We have $r \neq 0$ because $\operatorname{gcd}(a, b)=1$. Hence

$$
f(a, b)=f(r, b)=1-f(b, r), \quad \frac{a}{b}=\left[q ; a_{1}, \ldots, a_{k}\right], \quad \text { and } \quad \frac{b}{r}=\left[a_{1} ; a_{2}, \ldots, a_{k}\right] .
$$

Then the number of terms in the continued fraction representations of $a / b$ and $b / r$ differ by one. Since $r<b$, the inductive hypothesis yields

$$
f(b, r)=0 \Longleftrightarrow k-1 \text { is even, }
$$

and thus

$$
f(a, b)=0 \Longleftrightarrow f(b, r)=1 \Longleftrightarrow k-1 \text { is odd } \Longleftrightarrow k \text { is even. }
$$

Now we use the following well-known properties of continued fractions to prove the Lemma:
Let $p_{i}$ and $q_{i}$ be coprime positive integers with $\left[a_{0} ; a_{1}, a_{2}, \ldots, a_{i}\right]=p_{i} / q_{i}$, with the notation borrowed from the Claim. In particular, $a / b=\left[a_{0} ; a_{1}, a_{2}, \ldots, a_{k}\right]=p_{k} / q_{k}$. Assume that $k>0$ and define $q_{-1}=0$ if necessary. Then

- $q_{k}=a_{k} q_{k-1}+q_{k-2}$, and
- $a q_{k-1}-b p_{k-1}=p_{k} q_{k-1}-q_{k} p_{k-1}=(-1)^{k-1}$.

Assume that $k>0$. Then $a_{k} \geqslant 2$, and

$$
b=q_{k}=a_{k} q_{k-1}+q_{k-2} \geqslant a_{k} q_{k-1} \geqslant 2 q_{k-1} \Longrightarrow q_{k-1} \leqslant \frac{b}{2},
$$

with strict inequality for $k>1$, and

$$
(-1)^{k-1} q_{k-1} a+(-1)^{k} p_{k-1} b=1 .
$$

Now we finish the proof of the Lemma. It is immediate for $k=0$. If $k=1$, then $(-1)^{k-1}=1$, so

$$
-b / 2<0 \leqslant(-1)^{k-1} q_{k-1} \leqslant b / 2 .
$$

If $k>1$, we have $q_{k-1}<b / 2$, so

$$
-b / 2<(-1)^{k-1} q_{k-1}<b / 2 .
$$

Thus, for any $k>0$, we find that $g(a, b)=\left((-1)^{k-1} q_{k-1},(-1)^{k} p_{k-1}\right)$, and so

$$
f(a, b)=1 \Longleftrightarrow k \text { is odd } \Longleftrightarrow u=(-1)^{k-1} q_{k-1}>0 .
$$

Comment 1. The Lemma can also be established by observing that $f$ is uniquely defined on $\mathbb{A}$, defining $f_{1}(a, b)=1$ if $u>0$ in $g(a, b)=(u, v)$ and $f_{1}(a, b)=0$ otherwise, and verifying that $f_{1}$ satisfies all the conditions from the statement.

It seems that the main difficulty of the problem is in conjecturing the Lemma.
Comment 2. The case $p \equiv 1(\bmod 4)$ is, in fact, easier than the original problem. We have, in general, for $1 \leqslant a \leqslant p-1$,
$f(a, p)=1-f(p, a)=1-f(p-a, a)=f(a, p-a)=f(a+(p-a), p-a)=f(p, p-a)=1-f(p-a, p)$.
If $p \equiv 1(\bmod 4)$, then $a$ is a quadratic residue modulo $p$ if and only if $p-a$ is a quadratic residue modulo $p$. Therefore, denoting by $r_{k}$ (with $1 \leqslant r_{k} \leqslant p-1$ ) the remainder of the division of $k^{2}$ by $p$, we get

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right)=\sum_{n=1}^{p-1} f\left(r_{n}, p\right)=\frac{1}{2} \sum_{n=1}^{p-1}\left(f\left(r_{n}, p\right)+f\left(p-r_{n}, p\right)\right)=\frac{p-1}{2} .
$$

Comment 3. The estimate for the sum $\sum_{n=1}^{p} f\left(n^{2}, p\right)$ can be improved by refining the final argument in Solution 1. In fact, one can prove that

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right) \geqslant \frac{p-1}{16}
$$

By counting the number of perfect squares in the intervals $[k p,(k+1 / 2) p)$, we find that

$$
\begin{equation*}
\sum_{n=1}^{p-1} f\left(n^{2}, p\right)=\sum_{k=0}^{p-1}\left(\left\lfloor\sqrt{\left(k+\frac{1}{2}\right) p}\right\rfloor-\lfloor\sqrt{k p}\rfloor\right) . \tag{2}
\end{equation*}
$$

Each summand of (2) is non-negative. We now estimate the number of positive summands. Suppose that a summand is zero, i.e.,

$$
\left\lfloor\sqrt{\left(k+\frac{1}{2}\right) p}\right\rfloor=\lfloor\sqrt{k p}\rfloor=: q .
$$

Then both of the numbers $k p$ and $k p+p / 2$ lie within the interval $\left[q^{2},(q+1)^{2}\right)$. Hence

$$
\frac{p}{2}<(q+1)^{2}-q^{2}
$$

which implies

$$
q \geqslant \frac{p-1}{4} .
$$

Since $q \leqslant \sqrt{k p}$, if the $k^{\text {th }}$ summand of (2) is zero, then

$$
k \geqslant \frac{q^{2}}{p} \geqslant \frac{(p-1)^{2}}{16 p}>\frac{p-2}{16} \Longrightarrow k \geqslant \frac{p-1}{16} .
$$

So at least the first $\left\lceil\frac{p-1}{16}\right\rceil$ summands (from $k=0$ to $k=\left\lceil\frac{p-1}{16}\right\rceil-1$ ) are positive, and the result follows.

Comment 4. The bound can be further improved by using different methods. In fact, we prove that

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right) \geqslant \frac{p-3}{4}
$$

To that end, we use the Legendre symbol

$$
\left(\frac{a}{p}\right)= \begin{cases}0 & \text { if } p \mid a \\ 1 & \text { if } a \text { is a nonzero quadratic residue } \bmod p \\ -1 & \text { otherwise }\end{cases}
$$

We start with the following Claim, which tells us that there are not too many consecutive quadratic residues or consecutive quadratic non-residues.

Claim. $\sum_{n=1}^{p-1}\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=-1$.
Proof. We have $\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=\left(\frac{n(n+1)}{p}\right)$. For $1 \leqslant n \leqslant p-1$, we get that $n(n+1) \equiv n^{2}\left(1+n^{-1}\right)(\bmod p)$, hence $\left(\frac{n(n+1)}{p}\right)=\left(\frac{1+n^{-1}}{p}\right)$. Since $\left\{1+n^{-1} \bmod p: 1 \leqslant n \leqslant p-1\right\}=\{0,2,3, \ldots, p-1 \bmod p\}$, we find

$$
\sum_{n=1}^{p-1}\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=\sum_{n=1}^{p-1}\left(\frac{1+n^{-1}}{p}\right)=\sum_{n=1}^{p-1}\left(\frac{n}{p}\right)-1=-1
$$

because $\sum_{n=1}^{p}\left(\frac{n}{p}\right)=0$.
Observe that (1) becomes

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right)=2|S|, \quad S=\left\{r: 1 \leqslant r \leqslant \frac{p-1}{2} \text { and }\left(\frac{r}{p}\right)=1\right\} .
$$

We connect $S$ with the sum from the claim by pairing quadratic residues and quadratic non-residues. To that end, define

$$
\begin{aligned}
S^{\prime} & =\left\{r: 1 \leqslant r \leqslant \frac{p-1}{2} \text { and }\left(\frac{r}{p}\right)=-1\right\} \\
T & =\left\{r: \frac{p+1}{2} \leqslant r \leqslant p-1 \text { and }\left(\frac{r}{p}\right)=1\right\} \\
T^{\prime} & =\left\{r: \frac{p+1}{2} \leqslant r \leqslant p-1 \text { and }\left(\frac{r}{p}\right)=-1\right\}
\end{aligned}
$$

Since there are exactly $(p-1) / 2$ nonzero quadratic residues modulo $p,|S|+|T|=(p-1) / 2$. Also we obviously have $|T|+\left|T^{\prime}\right|=(p-1) / 2$. Then $|S|=\left|T^{\prime}\right|$.

For the sake of brevity, define $t=|S|=\left|T^{\prime}\right|$. If $\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=-1$, then exactly of one the numbers $\left(\frac{n}{p}\right)$ and $\left(\frac{n+1}{p}\right)$ is equal to 1 , so

$$
\left\lvert\,\left\{n: 1 \leqslant n \leqslant \frac{p-3}{2} \text { and }\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=-1\right\}|\leqslant|S|+|S-1|=2 t .\right.
$$

On the other hand, if $\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=-1$, then exactly one of $\left(\frac{n}{p}\right)$ and $\left(\frac{n+1}{p}\right)$ is equal to -1 , and

$$
\left\lvert\,\left\{n: \frac{p+1}{2} \leqslant n \leqslant p-2 \text { and }\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=-1\right\}\left|\leqslant\left|T^{\prime}\right|+\left|T^{\prime}-1\right|=2 t .\right.\right.
$$

Thus, taking into account that the middle term $\left(\frac{(p-1) / 2}{p}\right)\left(\frac{(p+1) / 2}{p}\right)$ may happen to be -1 ,

$$
\left.\left\lvert\,\left\{n: 1 \leqslant n \leqslant p-2 \text { and }\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=-1\right\}\right. \right\rvert\, \leqslant 4 t+1 .
$$

This implies that

$$
\left.\left\lvert\,\left\{n: 1 \leqslant n \leqslant p-2 \text { and }\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right)=1\right\}\right. \right\rvert\, \geqslant(p-2)-(4 t+1)=p-4 t-3,
$$

and so

$$
-1=\sum_{n=1}^{p-1}\left(\frac{n}{p}\right)\left(\frac{n+1}{p}\right) \geqslant p-4 t-3-(4 t+1)=p-8 t-4,
$$

which implies $8 t \geqslant p-3$, and thus

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right)=2 t \geqslant \frac{p-3}{4}
$$

Comment 5. It is possible to prove that

$$
\sum_{n=1}^{p-1} f\left(n^{2}, p\right) \geqslant \frac{p-1}{2}
$$

The case $p \equiv 1(\bmod 4)$ was already mentioned, and it is the equality case. If $p \equiv 3(\bmod 4)$, then, by a theorem of Dirichlet, we have

$$
\left.\left\lvert\,\left\{r: 1 \leqslant r \leqslant \frac{p-1}{2} \text { and }\left(\frac{r}{p}\right)=1\right\}\right. \right\rvert\,>\frac{p-1}{4},
$$

which implies the result.
See https://en.wikipedia.org/wiki/Quadratic_residue\#Dirichlet.27s_formulas for the full statement of the theorem. It seems that no elementary proof of it is known; a proof using complex analysis is available, for instance, in Chapter 7 of the book Quadratic Residues and Non-Residues: Selected Topics, by Steve Wright, available in https://arxiv.org/abs/1408.0235.

BIÊNIO DA
MATEMATICA BRASIL

## SHORTLISTED PROBLEMS

## WITH SOLUTIONS




# Shortlisted Problems (with solutions) 

$59^{\text {th }}$ International Mathematical Olympiad
Cluj-Napoca — Romania, 3-14 July 2018

# The Shortlist has to be kept strictly confidential until the conclusion of the following International Mathematical Olympiad. IMO General Regulations $\S 6.6$ 

## Contributing Countries

The Organising Committee and the Problem Selection Committee of IMO 2018 thank the following 49 countries for contributing 168 problem proposals:

Armenia, Australia, Austria, Azerbaijan, Belarus, Belgium, Bosnia and Herzegovina, Brazil, Bulgaria, Canada, China, Croatia, Cyprus, Czech Republic, Denmark, Estonia, Germany, Greece, Hong Kong, Iceland, India, Indonesia, Iran, Ireland, Israel, Japan, Kosovo, Luxembourg, Mexico, Moldova, Mongolia, Netherlands, Nicaragua, Poland, Russia, Serbia, Singapore, Slovakia, Slovenia, South Africa, South Korea, Switzerland, Taiwan, Tanzania, Thailand, Turkey, Ukraine, United Kingdom, U.S.A.

## Problem Selection Committee



Calin Popescu, Radu Gologan, Marian Andronache, Mihail Baluna, Nicolae Beli, Ilya Bogdanov, Pavel Kozhevnikov, Géza Kós, Sever Moldoveanu

## Problems

## Algebra

A1. Let $\mathbb{Q}_{>0}$ denote the set of all positive rational numbers. Determine all functions $f: \mathbb{Q}_{>0} \rightarrow \mathbb{Q}_{>0}$ satisfying

$$
f\left(x^{2} f(y)^{2}\right)=f(x)^{2} f(y)
$$

for all $x, y \in \mathbb{Q}_{>0}$.
(Switzerland)
A2. Find all positive integers $n \geqslant 3$ for which there exist real numbers $a_{1}, a_{2}, \ldots, a_{n}$, $a_{n+1}=a_{1}, a_{n+2}=a_{2}$ such that

$$
a_{i} a_{i+1}+1=a_{i+2}
$$

for all $i=1,2, \ldots, n$.
(Slovakia)
A3. Given any set $S$ of positive integers, show that at least one of the following two assertions holds:
(1) There exist distinct finite subsets $F$ and $G$ of $S$ such that $\sum_{x \in F} 1 / x=\sum_{x \in G} 1 / x$;
(2) There exists a positive rational number $r<1$ such that $\sum_{x \in F} 1 / x \neq r$ for all finite subsets $F$ of $S$.

A4. Let $a_{0}, a_{1}, a_{2}, \ldots$ be a sequence of real numbers such that $a_{0}=0, a_{1}=1$, and for every $n \geqslant 2$ there exists $1 \leqslant k \leqslant n$ satisfying

$$
a_{n}=\frac{a_{n-1}+\cdots+a_{n-k}}{k} .
$$

Find the maximal possible value of $a_{2018}-a_{2017}$.
(Belgium)
A5. Determine all functions $f:(0, \infty) \rightarrow \mathbb{R}$ satisfying

$$
\left(x+\frac{1}{x}\right) f(y)=f(x y)+f\left(\frac{y}{x}\right)
$$

for all $x, y>0$.
(South Korea)
A6. Let $m, n \geqslant 2$ be integers. Let $f\left(x_{1}, \ldots, x_{n}\right)$ be a polynomial with real coefficients such that

$$
f\left(x_{1}, \ldots, x_{n}\right)=\left\lfloor\frac{x_{1}+\ldots+x_{n}}{m}\right\rfloor \text { for every } x_{1}, \ldots, x_{n} \in\{0,1, \ldots, m-1\} .
$$

Prove that the total degree of $f$ is at least $n$.
(Brazil)
A7. Find the maximal value of

$$
S=\sqrt[3]{\frac{a}{b+7}}+\sqrt[3]{\frac{b}{c+7}}+\sqrt[3]{\frac{c}{d+7}}+\sqrt[3]{\frac{d}{a+7}}
$$

where $a, b, c, d$ are nonnegative real numbers which satisfy $a+b+c+d=100$.

## Combinatorics

C1. Let $n \geqslant 3$ be an integer. Prove that there exists a set $S$ of $2 n$ positive integers satisfying the following property: For every $m=2,3, \ldots, n$ the set $S$ can be partitioned into two subsets with equal sums of elements, with one of subsets of cardinality $m$.
(Iceland)
C2. Queenie and Horst play a game on a $20 \times 20$ chessboard. In the beginning the board is empty. In every turn, Horst places a black knight on an empty square in such a way that his new knight does not attack any previous knights. Then Queenie places a white queen on an empty square. The game gets finished when somebody cannot move.

Find the maximal positive $K$ such that, regardless of the strategy of Queenie, Horst can put at least $K$ knights on the board.
(Armenia)
C3. Let $n$ be a given positive integer. Sisyphus performs a sequence of turns on a board consisting of $n+1$ squares in a row, numbered 0 to $n$ from left to right. Initially, $n$ stones are put into square 0 , and the other squares are empty. At every turn, Sisyphus chooses any nonempty square, say with $k$ stones, takes one of those stones and moves it to the right by at most $k$ squares (the stone should stay within the board). Sisyphus' aim is to move all $n$ stones to square $n$.

Prove that Sisyphus cannot reach the aim in less than

$$
\left\lceil\frac{n}{1}\right\rceil+\left\lceil\frac{n}{2}\right\rceil+\left\lceil\frac{n}{3}\right\rceil+\cdots+\left\lceil\frac{n}{n}\right\rceil
$$

turns. (As usual, $\lceil x\rceil$ stands for the least integer not smaller than $x$.)
(Netherlands)
C4. An anti-Pascal pyramid is a finite set of numbers, placed in a triangle-shaped array so that the first row of the array contains one number, the second row contains two numbers, the third row contains three numbers and so on; and, except for the numbers in the bottom row, each number equals the absolute value of the difference of the two numbers below it. For instance, the triangle below is an anti-Pascal pyramid with four rows, in which every integer from 1 to $1+2+3+4=10$ occurs exactly once:

$$
\begin{aligned}
& 4 \\
& 26 \\
& \begin{array}{lll}
5 & 7 & 1
\end{array} \\
& 8 \quad 3 \quad 10 \quad 9 .
\end{aligned}
$$

Is it possible to form an anti-Pascal pyramid with 2018 rows, using every integer from 1 to $1+2+\cdots+2018$ exactly once?
(Iran)
C5. Let $k$ be a positive integer. The organising committee of a tennis tournament is to schedule the matches for $2 k$ players so that every two players play once, each day exactly one match is played, and each player arrives to the tournament site the day of his first match, and departs the day of his last match. For every day a player is present on the tournament, the committee has to pay 1 coin to the hotel. The organisers want to design the schedule so as to minimise the total cost of all players' stays. Determine this minimum cost.

C6. Let $a$ and $b$ be distinct positive integers. The following infinite process takes place on an initially empty board.
( $i$ If there is at least a pair of equal numbers on the board, we choose such a pair and increase one of its components by $a$ and the other by $b$.
(ii) If no such pair exists, we write down two times the number 0 .

Prove that, no matter how we make the choices in (i), operation (ii) will be performed only finitely many times.
(Serbia)
C7. Consider 2018 pairwise crossing circles no three of which are concurrent. These circles subdivide the plane into regions bounded by circular edges that meet at vertices. Notice that there are an even number of vertices on each circle. Given the circle, alternately colour the vertices on that circle red and blue. In doing so for each circle, every vertex is coloured twice once for each of the two circles that cross at that point. If the two colourings agree at a vertex, then it is assigned that colour; otherwise, it becomes yellow. Show that, if some circle contains at least 2061 yellow points, then the vertices of some region are all yellow.

## Geometry

G1. Let $A B C$ be an acute-angled triangle with circumcircle $\Gamma$. Let $D$ and $E$ be points on the segments $A B$ and $A C$, respectively, such that $A D=A E$. The perpendicular bisectors of the segments $B D$ and $C E$ intersect the small arcs $\overline{A B}$ and $\overparen{A C}$ at points $F$ and $G$ respectively. Prove that $D E \| F G$.
(Greece)
G2. Let $A B C$ be a triangle with $A B=A C$, and let $M$ be the midpoint of $B C$. Let $P$ be a point such that $P B<P C$ and $P A$ is parallel to $B C$. Let $X$ and $Y$ be points on the lines $P B$ and $P C$, respectively, so that $B$ lies on the segment $P X, C$ lies on the segment $P Y$, and $\angle P X M=\angle P Y M$. Prove that the quadrilateral $A P X Y$ is cyclic.
(Australia)
G3. A circle $\omega$ of radius 1 is given. A collection $T$ of triangles is called good, if the following conditions hold:
(i) each triangle from $T$ is inscribed in $\omega$;
(ii) no two triangles from $T$ have a common interior point.

Determine all positive real numbers $t$ such that, for each positive integer $n$, there exists a good collection of $n$ triangles, each of perimeter greater than $t$.
(South Africa)
G4. A point $T$ is chosen inside a triangle $A B C$. Let $A_{1}, B_{1}$, and $C_{1}$ be the reflections of $T$ in $B C, C A$, and $A B$, respectively. Let $\Omega$ be the circumcircle of the triangle $A_{1} B_{1} C_{1}$. The lines $A_{1} T, B_{1} T$, and $C_{1} T$ meet $\Omega$ again at $A_{2}, B_{2}$, and $C_{2}$, respectively. Prove that the lines $A A_{2}, B B_{2}$, and $C C_{2}$ are concurrent on $\Omega$.
(Mongolia)
G5. Let $A B C$ be a triangle with circumcircle $\omega$ and incentre $I$. A line $\ell$ intersects the lines $A I, B I$, and $C I$ at points $D, E$, and $F$, respectively, distinct from the points $A, B, C$, and $I$. The perpendicular bisectors $x, y$, and $z$ of the segments $A D, B E$, and $C F$, respectively determine a triangle $\Theta$. Show that the circumcircle of the triangle $\Theta$ is tangent to $\omega$.
(Denmark)
G6. A convex quadrilateral $A B C D$ satisfies $A B \cdot C D=B C \cdot D A$. A point $X$ is chosen inside the quadrilateral so that $\angle X A B=\angle X C D$ and $\angle X B C=\angle X D A$. Prove that $\angle A X B+$ $\angle C X D=180^{\circ}$.
(Poland)
G7. Let $O$ be the circumcentre, and $\Omega$ be the circumcircle of an acute-angled triangle $A B C$. Let $P$ be an arbitrary point on $\Omega$, distinct from $A, B, C$, and their antipodes in $\Omega$. Denote the circumcentres of the triangles $A O P, B O P$, and $C O P$ by $O_{A}, O_{B}$, and $O_{C}$, respectively. The lines $\ell_{A}, \ell_{B}$, and $\ell_{C}$ perpendicular to $B C, C A$, and $A B$ pass through $O_{A}, O_{B}$, and $O_{C}$, respectively. Prove that the circumcircle of the triangle formed by $\ell_{A}, \ell_{B}$, and $\ell_{C}$ is tangent to the line $O P$.

## Number Theory

N1. Determine all pairs $(n, k)$ of distinct positive integers such that there exists a positive integer $s$ for which the numbers of divisors of $s n$ and of $s k$ are equal.
(Ukraine)
N2. Let $n>1$ be a positive integer. Each cell of an $n \times n$ table contains an integer. Suppose that the following conditions are satisfied:
(i) Each number in the table is congruent to 1 modulo $n$;
(ii) The sum of numbers in any row, as well as the sum of numbers in any column, is congruent to $n$ modulo $n^{2}$.

Let $R_{i}$ be the product of the numbers in the $i^{\text {th }}$ row, and $C_{j}$ be the product of the numbers in the $j^{\text {th }}$ column. Prove that the sums $R_{1}+\cdots+R_{n}$ and $C_{1}+\cdots+C_{n}$ are congruent modulo $n^{4}$.
(Indonesia)
N3. Define the sequence $a_{0}, a_{1}, a_{2}, \ldots$ by $a_{n}=2^{n}+2^{\lfloor n / 2\rfloor}$. Prove that there are infinitely many terms of the sequence which can be expressed as a sum of (two or more) distinct terms of the sequence, as well as infinitely many of those which cannot be expressed in such a way.
(Serbia)
N4. Let $a_{1}, a_{2}, \ldots, a_{n}, \ldots$ be a sequence of positive integers such that

$$
\frac{a_{1}}{a_{2}}+\frac{a_{2}}{a_{3}}+\cdots+\frac{a_{n-1}}{a_{n}}+\frac{a_{n}}{a_{1}}
$$

is an integer for all $n \geqslant k$, where $k$ is some positive integer. Prove that there exists a positive integer $m$ such that $a_{n}=a_{n+1}$ for all $n \geqslant m$.
(Mongolia)
N5. Four positive integers $x, y, z$, and $t$ satisfy the relations

$$
x y-z t=x+y=z+t .
$$

Is it possible that both $x y$ and $z t$ are perfect squares?
(Russia)
N6. Let $f:\{1,2,3, \ldots\} \rightarrow\{2,3, \ldots\}$ be a function such that $f(m+n) \mid f(m)+f(n)$ for all pairs $m, n$ of positive integers. Prove that there exists a positive integer $c>1$ which divides all values of $f$.
(Mexico)
Let $n \geqslant 2018$ be an integer, and let $a_{1}, a_{2}, \ldots, a_{n}, b_{1}, b_{2}, \ldots, b_{n}$ be pairwise distinct positive integers not exceeding $5 n$. Suppose that the sequence

$$
\frac{a_{1}}{b_{1}}, \frac{a_{2}}{b_{2}}, \ldots, \frac{a_{n}}{b_{n}}
$$

forms an arithmetic progression. Prove that the terms of the sequence are equal.

## Solutions

## Algebra

A1. Let $\mathbb{Q}_{>0}$ denote the set of all positive rational numbers. Determine all functions $f: \mathbb{Q}_{>0} \rightarrow \mathbb{Q}_{>0}$ satisfying

$$
\begin{equation*}
f\left(x^{2} f(y)^{2}\right)=f(x)^{2} f(y) \tag{*}
\end{equation*}
$$

for all $x, y \in \mathbb{Q}_{>0}$.
(Switzerland)
Answer: $f(x)=1$ for all $x \in \mathbb{Q}_{>0}$.
Solution. Take any $a, b \in \mathbb{Q}_{>0}$. By substituting $x=f(a), y=b$ and $x=f(b), y=a$ into (*) we get

$$
f(f(a))^{2} f(b)=f\left(f(a)^{2} f(b)^{2}\right)=f(f(b))^{2} f(a)
$$

which yields

$$
\frac{f(f(a))^{2}}{f(a)}=\frac{f(f(b))^{2}}{f(b)} \quad \text { for all } a, b \in \mathbb{Q}_{>0} .
$$

In other words, this shows that there exists a constant $C \in \mathbb{Q}_{>0}$ such that $f(f(a))^{2}=C f(a)$, or

$$
\begin{equation*}
\left(\frac{f(f(a))}{C}\right)^{2}=\frac{f(a)}{C} \quad \text { for all } a \in \mathbb{Q}_{>0} \tag{1}
\end{equation*}
$$



$$
\frac{f(a)}{C}=\left(\frac{f^{2}(a)}{C}\right)^{2}=\left(\frac{f^{3}(a)}{C}\right)^{4}=\cdots=\left(\frac{f^{n+1}(a)}{C}\right)^{2^{n}}
$$

for all positive integer $n$. So, $f(a) / C$ is the $2^{n}$-th power of a rational number for all positive integer $n$. This is impossible unless $f(a) / C=1$, since otherwise the exponent of some prime in the prime decomposition of $f(a) / C$ is not divisible by sufficiently large powers of 2 . Therefore, $f(a)=C$ for all $a \in \mathbb{Q}_{>0}$.

Finally, after substituting $f \equiv C$ into (*) we get $C=C^{3}$, whence $C=1$. So $f(x) \equiv 1$ is the unique function satisfying (*).

Comment 1. There are several variations of the solution above. For instance, one may start with finding $f(1)=1$. To do this, let $d=f(1)$. By substituting $x=y=1$ and $x=d^{2}, y=1$ into (*) we get $f\left(d^{2}\right)=d^{3}$ and $f\left(d^{6}\right)=f\left(d^{2}\right)^{2} \cdot d=d^{7}$. By substituting now $x=1, y=d^{2}$ we obtain $f\left(d^{6}\right)=d^{2} \cdot d^{3}=d^{5}$. Therefore, $d^{7}=f\left(d^{6}\right)=d^{5}$, whence $d=1$.

After that, the rest of the solution simplifies a bit, since we already know that $C=\frac{f(f(1))^{2}}{f(1)}=1$. Hence equation (1) becomes merely $f(f(a))^{2}=f(a)$, which yields $f(a)=1$ in a similar manner.

Comment 2. There exist nonconstant functions $f: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$satisfying (*) for all real $x, y>0-$ e.g., $f(x)=\sqrt{x}$.

A2. Find all positive integers $n \geqslant 3$ for which there exist real numbers $a_{1}, a_{2}, \ldots, a_{n}$, $a_{n+1}=a_{1}, a_{n+2}=a_{2}$ such that

$$
a_{i} a_{i+1}+1=a_{i+2}
$$

for all $i=1,2, \ldots, n$.
(Slovakia)
Answer: $n$ can be any multiple of 3 .
Solution 1. For the sake of convenience, extend the sequence $a_{1}, \ldots, a_{n+2}$ to an infinite periodic sequence with period $n$. ( $n$ is not necessarily the shortest period.)

If $n$ is divisible by 3 , then $\left(a_{1}, a_{2}, \ldots\right)=(-1,-1,2,-1,-1,2, \ldots)$ is an obvious solution.
We will show that in every periodic sequence satisfying the recurrence, each positive term is followed by two negative values, and after them the next number is positive again. From this, it follows that $n$ is divisible by 3 .

If the sequence contains two consecutive positive numbers $a_{i}, a_{i+1}$, then $a_{i+2}=a_{i} a_{i+1}+1>1$, so the next value is positive as well; by induction, all numbers are positive and greater than 1 . But then $a_{i+2}=a_{i} a_{i+1}+1 \geqslant 1 \cdot a_{i+1}+1>a_{i+1}$ for every index $i$, which is impossible: our sequence is periodic, so it cannot increase everywhere.

If the number 0 occurs in the sequence, $a_{i}=0$ for some index $i$, then it follows that $a_{i+1}=a_{i-1} a_{i}+1$ and $a_{i+2}=a_{i} a_{i+1}+1$ are two consecutive positive elements in the sequences and we get the same contradiction again.

Notice that after any two consecutive negative numbers the next one must be positive: if $a_{i}<0$ and $a_{i+1}<0$, then $a_{i+2}=a_{1} a_{i+1}+1>1>0$. Hence, the positive and negative numbers follow each other in such a way that each positive term is followed by one or two negative values and then comes the next positive term.

Consider the case when the positive and negative values alternate. So, if $a_{i}$ is a negative value then $a_{i+1}$ is positive, $a_{i+2}$ is negative and $a_{i+3}$ is positive again.

Notice that $a_{i} a_{i+1}+1=a_{i+2}<0<a_{i+3}=a_{i+1} a_{i+2}+1$; by $a_{i+1}>0$ we conclude $a_{i}<a_{i+2}$. Hence, the negative values form an infinite increasing subsequence, $a_{i}<a_{i+2}<a_{i+4}<\ldots$, which is not possible, because the sequence is periodic.

The only case left is when there are consecutive negative numbers in the sequence. Suppose that $a_{i}$ and $a_{i+1}$ are negative; then $a_{i+2}=a_{i} a_{i+1}+1>1$. The number $a_{i+3}$ must be negative. We show that $a_{i+4}$ also must be negative.

Notice that $a_{i+3}$ is negative and $a_{i+4}=a_{i+2} a_{i+3}+1<1<a_{i} a_{i+1}+1=a_{i+2}$, so

$$
a_{i+5}-a_{i+4}=\left(a_{i+3} a_{i+4}+1\right)-\left(a_{i+2} a_{i+3}+1\right)=a_{i+3}\left(a_{i+4}-a_{i+2}\right)>0,
$$

therefore $a_{i+5}>a_{i+4}$. Since at most one of $a_{i+4}$ and $a_{i+5}$ can be positive, that means that $a_{i+4}$ must be negative.

Now $a_{i+3}$ and $a_{i+4}$ are negative and $a_{i+5}$ is positive; so after two negative and a positive terms, the next three terms repeat the same pattern. That completes the solution.

Solution 2. We prove that the shortest period of the sequence must be 3. Then it follows that $n$ must be divisible by 3 .

Notice that the equation $x^{2}+1=x$ has no real root, so the numbers $a_{1}, \ldots, a_{n}$ cannot be all equal, hence the shortest period of the sequence cannot be 1 .

By applying the recurrence relation for $i$ and $i+1$,

$$
\begin{gathered}
\left(a_{i+2}-1\right) a_{i+2}=a_{i} a_{i+1} a_{i+2}=a_{i}\left(a_{i+3}-1\right), \quad \text { so } \\
a_{i+2}^{2}-a_{i} a_{i+3}=a_{i+2}-a_{i} .
\end{gathered}
$$

By summing over $i=1,2, \ldots, n$, we get

$$
\sum_{i=1}^{n}\left(a_{i}-a_{i+3}\right)^{2}=0
$$

That proves that $a_{i}=a_{i+3}$ for every index $i$, so the sequence $a_{1}, a_{2}, \ldots$ is indeed periodic with period 3. The shortest period cannot be 1 , so it must be 3 ; therefore, $n$ is divisible by 3 .

Comment. By solving the system of equations $a b+1=c, \quad b c+1=a, \quad c a+1=b$, it can be seen that the pattern $(-1,-1,2)$ is repeated in all sequences satisfying the problem conditions.

A3. Given any set $S$ of positive integers, show that at least one of the following two assertions holds:
(1) There exist distinct finite subsets $F$ and $G$ of $S$ such that $\sum_{x \in F} 1 / x=\sum_{x \in G} 1 / x$;
(2) There exists a positive rational number $r<1$ such that $\sum_{x \in F} 1 / x \neq r$ for all finite subsets $F$ of $S$.

Solution 1. Argue indirectly. Agree, as usual, that the empty sum is 0 to consider rationals in $\left[0,1\right.$ ); adjoining 0 causes no harm, since $\sum_{x \in F} 1 / x=0$ for no nonempty finite subset $F$ of $S$. For every rational $r$ in $[0,1)$, let $F_{r}$ be the unique finite subset of $S$ such that $\sum_{x \in F_{r}} 1 / x=r$. The argument hinges on the lemma below.
Lemma. If $x$ is a member of $S$ and $q$ and $r$ are rationals in $[0,1)$ such that $q-r=1 / x$, then $x$ is a member of $F_{q}$ if and only if it is not one of $F_{r}$.
Proof. If $x$ is a member of $F_{q}$, then

$$
\sum_{y \in F_{q} \backslash\{x\}} \frac{1}{y}=\sum_{y \in F_{q}} \frac{1}{y}-\frac{1}{x}=q-\frac{1}{x}=r=\sum_{y \in F_{r}} \frac{1}{y},
$$

so $F_{r}=F_{q} \backslash\{x\}$, and $x$ is not a member of $F_{r}$. Conversely, if $x$ is not a member of $F_{r}$, then

$$
\sum_{y \in F_{r} \cup\{x\}} \frac{1}{y}=\sum_{y \in F_{r}} \frac{1}{y}+\frac{1}{x}=r+\frac{1}{x}=q=\sum_{y \in F_{q}} \frac{1}{y},
$$

so $F_{q}=F_{r} \cup\{x\}$, and $x$ is a member of $F_{q}$.
Consider now an element $x$ of $S$ and a positive rational $r<1$. Let $n=\lfloor r x\rfloor$ and consider the sets $F_{r-k / x}, k=0, \ldots, n$. Since $0 \leqslant r-n / x<1 / x$, the set $F_{r-n / x}$ does not contain $x$, and a repeated application of the lemma shows that the $F_{r-(n-2 k) / x}$ do not contain $x$, whereas the $F_{r-(n-2 k-1) / x}$ do. Consequently, $x$ is a member of $F_{r}$ if and only if $n$ is odd.

Finally, consider $F_{2 / 3}$. By the preceding, $\lfloor 2 x / 3\rfloor$ is odd for each $x$ in $F_{2 / 3}$, so $2 x / 3$ is not integral. Since $F_{2 / 3}$ is finite, there exists a positive rational $\varepsilon$ such that $\lfloor(2 / 3-\varepsilon) x\rfloor=\lfloor 2 x / 3\rfloor$ for all $x$ in $F_{2 / 3}$. This implies that $F_{2 / 3}$ is a subset of $F_{2 / 3-\varepsilon}$ which is impossible.

Comment. The solution above can be adapted to show that the problem statement still holds, if the condition $r<1$ in (2) is replaced with $r<\delta$, for an arbitrary positive $\delta$. This yields that, if $S$ does not satisfy (1), then there exist infinitely many positive rational numbers $r<1$ such that $\sum_{x \in F} 1 / x \neq r$ for all finite subsets $F$ of $S$.

Solution 2. A finite $S$ clearly satisfies (2), so let $S$ be infinite. If $S$ fails both conditions, so does $S \backslash\{1\}$. We may and will therefore assume that $S$ consists of integers greater than 1 . Label the elements of $S$ increasingly $x_{1}<x_{2}<\cdots$, where $x_{1} \geqslant 2$.

We first show that $S$ satisfies (2) if $x_{n+1} \geqslant 2 x_{n}$ for all $n$. In this case, $x_{n} \geqslant 2^{n-1} x_{1}$ for all $n$, so

$$
s=\sum_{n \geqslant 1} \frac{1}{x_{n}} \leqslant \sum_{n \geqslant 1} \frac{1}{2^{n-1} x_{1}}=\frac{2}{x_{1}} .
$$

If $x_{1} \geqslant 3$, or $x_{1}=2$ and $x_{n+1}>2 x_{n}$ for some $n$, then $\sum_{x \in F} 1 / x<s<1$ for every finite subset $F$ of $S$, so $S$ satisfies (2); and if $x_{1}=2$ and $x_{n+1}=2 x_{n}$ for all $n$, that is, $x_{n}=2^{n}$ for all $n$, then every finite subset $F$ of $S$ consists of powers of 2 , so $\sum_{x \in F} 1 / x \neq 1 / 3$ and again $S$ satisfies (2).

Finally, we deal with the case where $x_{n+1}<2 x_{n}$ for some $n$. Consider the positive rational $r=1 / x_{n}-1 / x_{n+1}<1 / x_{n+1}$. If $r=\sum_{x \in F} 1 / x$ for no finite subset $F$ of $S$, then $S$ satisfies (2).

We now assume that $r=\sum_{x \in F_{0}} 1 / x$ for some finite subset $F_{0}$ of $S$, and show that $S$ satisfies (1). Since $\sum_{x \in F_{0}} 1 / x=r<1 / x_{n+1}$, it follows that $x_{n+1}$ is not a member of $F_{0}$, so

$$
\sum_{x \in F_{0} \cup\left\{x_{n+1}\right\}} \frac{1}{x}=\sum_{x \in F_{0}} \frac{1}{x}+\frac{1}{x_{n+1}}=r+\frac{1}{x_{n+1}}=\frac{1}{x_{n}} .
$$

Consequently, $F=F_{0} \cup\left\{x_{n+1}\right\}$ and $G=\left\{x_{n}\right\}$ are distinct finite subsets of $S$ such that $\sum_{x \in F} 1 / x=\sum_{x \in G} 1 / x$, and $S$ satisfies (1).

A4. Let $a_{0}, a_{1}, a_{2}, \ldots$ be a sequence of real numbers such that $a_{0}=0, a_{1}=1$, and for every $n \geqslant 2$ there exists $1 \leqslant k \leqslant n$ satisfying

$$
a_{n}=\frac{a_{n-1}+\cdots+a_{n-k}}{k} .
$$

Find the maximal possible value of $a_{2018}-a_{2017}$.
(Belgium)
Answer: The maximal value is $\frac{2016}{2017^{2}}$.
Solution 1. The claimed maximal value is achieved at

$$
\begin{gathered}
a_{1}=a_{2}=\cdots=a_{2016}=1, \quad a_{2017}=\frac{a_{2016}+\cdots+a_{0}}{2017}=1-\frac{1}{2017}, \\
a_{2018}=\frac{a_{2017}+\cdots+a_{1}}{2017}=1-\frac{1}{2017^{2}} .
\end{gathered}
$$

Now we need to show that this value is optimal. For brevity, we use the notation

$$
S(n, k)=a_{n-1}+a_{n-2}+\cdots+a_{n-k} \quad \text { for nonnegative integers } k \leqslant n .
$$

In particular, $S(n, 0)=0$ and $S(n, 1)=a_{n-1}$. In these terms, for every integer $n \geqslant 2$ there exists a positive integer $k \leqslant n$ such that $a_{n}=S(n, k) / k$.

For every integer $n \geqslant 1$ we define

$$
M_{n}=\max _{1 \leqslant k \leqslant n} \frac{S(n, k)}{k}, \quad m_{n}=\min _{1 \leqslant k \leqslant n} \frac{S(n, k)}{k}, \quad \text { and } \quad \Delta_{n}=M_{n}-m_{n} \geqslant 0 .
$$

By definition, $a_{n} \in\left[m_{n}, M_{n}\right]$ for all $n \geqslant 2$; on the other hand, $a_{n-1}=S(n, 1) / 1 \in\left[m_{n}, M_{n}\right]$. Therefore,

$$
a_{2018}-a_{2017} \leqslant M_{2018}-m_{2018}=\Delta_{2018}
$$

and we are interested in an upper bound for $\Delta_{2018}$.
Also by definition, for any $0<k \leqslant n$ we have $k m_{n} \leqslant S(n, k) \leqslant k M_{n}$; notice that these inequalities are also valid for $k=0$.
Claim 1. For every $n>2$, we have $\Delta_{n} \leqslant \frac{n-1}{n} \Delta_{n-1}$.
Proof. Choose positive integers $k, \ell \leqslant n$ such that $M_{n}=S(n, k) / k$ and $m_{n}=S(n, \ell) / \ell$. We have $S(n, k)=a_{n-1}+S(n-1, k-1)$, so

$$
k\left(M_{n}-a_{n-1}\right)=S(n, k)-k a_{n-1}=S(n-1, k-1)-(k-1) a_{n-1} \leqslant(k-1)\left(M_{n-1}-a_{n-1}\right),
$$

since $S(n-1, k-1) \leqslant(k-1) M_{n-1}$. Similarly, we get

$$
\ell\left(a_{n-1}-m_{n}\right)=(\ell-1) a_{n-1}-S(n-1, \ell-1) \leqslant(\ell-1)\left(a_{n-1}-m_{n-1}\right) .
$$

Since $m_{n-1} \leqslant a_{n-1} \leqslant M_{n-1}$ and $k, \ell \leqslant n$, the obtained inequalities yield

$$
\begin{aligned}
& M_{n}-a_{n-1} \leqslant \frac{k-1}{k}\left(M_{n-1}-a_{n-1}\right) \leqslant \frac{n-1}{n}\left(M_{n-1}-a_{n-1}\right) \quad \text { and } \\
& a_{n-1}-m_{n} \leqslant \frac{\ell-1}{\ell}\left(a_{n-1}-m_{n-1}\right) \leqslant \frac{n-1}{n}\left(a_{n-1}-m_{n-1}\right) .
\end{aligned}
$$

Therefore,

$$
\Delta_{n}=\left(M_{n}-a_{n-1}\right)+\left(a_{n-1}-m_{n}\right) \leqslant \frac{n-1}{n}\left(\left(M_{n-1}-a_{n-1}\right)+\left(a_{n-1}-m_{n-1}\right)\right)=\frac{n-1}{n} \Delta_{n-1}
$$

Back to the problem, if $a_{n}=1$ for all $n \leqslant 2017$, then $a_{2018} \leqslant 1$ and hence $a_{2018}-a_{2017} \leqslant 0$. Otherwise, let $2 \leqslant q \leqslant 2017$ be the minimal index with $a_{q}<1$. We have $S(q, i)=i$ for all $i=1,2, \ldots, q-1$, while $S(q, q)=q-1$. Therefore, $a_{q}<1$ yields $a_{q}=S(q, q) / q=1-\frac{1}{q}$.

Now we have $S(q+1, i)=i-\frac{1}{q}$ for $i=1,2, \ldots, q$, and $S(q+1, q+1)=q-\frac{1}{q}$. This gives us

$$
m_{q+1}=\frac{S(q+1,1)}{1}=\frac{S(q+1, q+1)}{q+1}=\frac{q-1}{q} \quad \text { and } \quad M_{q+1}=\frac{S(q+1, q)}{q}=\frac{q^{2}-1}{q^{2}}
$$

so $\Delta_{q+1}=M_{q+1}-m_{q+1}=(q-1) / q^{2}$. Denoting $N=2017 \geqslant q$ and using Claim 1 for $n=q+2, q+3, \ldots, N+1$ we finally obtain

$$
\Delta_{N+1} \leqslant \frac{q-1}{q^{2}} \cdot \frac{q+1}{q+2} \cdot \frac{q+2}{q+3} \cdots \frac{N}{N+1}=\frac{1}{N+1}\left(1-\frac{1}{q^{2}}\right) \leqslant \frac{1}{N+1}\left(1-\frac{1}{N^{2}}\right)=\frac{N-1}{N^{2}}
$$

as required.

Comment 1. One may check that the maximal value of $a_{2018}-a_{2017}$ is attained at the unique sequence, which is presented in the solution above.

Comment 2. An easier question would be to determine the maximal value of $\left|a_{2018}-a_{2017}\right|$. In this version, the answer $\frac{1}{2018}$ is achieved at

$$
a_{1}=a_{2}=\cdots=a_{2017}=1, \quad a_{2018}=\frac{a_{2017}+\cdots+a_{0}}{2018}=1-\frac{1}{2018} .
$$

To prove that this value is optimal, it suffices to notice that $\Delta_{2}=\frac{1}{2}$ and to apply Claim 1 obtaining

$$
\left|a_{2018}-a_{2017}\right| \leqslant \Delta_{2018} \leqslant \frac{1}{2} \cdot \frac{2}{3} \cdots \frac{2017}{2018}=\frac{1}{2018} .
$$

Solution 2. We present a different proof of the estimate $a_{2018}-a_{2017} \leqslant \frac{2016}{2017^{2}}$. We keep the same notations of $S(n, k), m_{n}$ and $M_{n}$ from the previous solution.

Notice that $S(n, n)=S(n, n-1)$, as $a_{0}=0$. Also notice that for $0 \leqslant k \leqslant \ell \leqslant n$ we have $S(n, \ell)=S(n, k)+S(n-k, \ell-k)$.
Claim 2. For every positive integer $n$, we have $m_{n} \leqslant m_{n+1}$ and $M_{n+1} \leqslant M_{n}$, so the segment [ $m_{n+1}, M_{n+1}$ ] is contained in [ $m_{n}, M_{n}$ ].
Proof. Choose a positive integer $k \leqslant n+1$ such that $m_{n+1}=S(n+1, k) / k$. Then we have

$$
k m_{n+1}=S(n+1, k)=a_{n}+S(n, k-1) \geqslant m_{n}+(k-1) m_{n}=k m_{n},
$$

which establishes the first inequality in the Claim. The proof of the second inequality is similar.

Claim 3. For every positive integers $k \geqslant n$, we have $m_{n} \leqslant a_{k} \leqslant M_{n}$.
Proof. By Claim 2, we have $\left[m_{k}, M_{k}\right] \subseteq\left[m_{k-1}, M_{k-1}\right] \subseteq \cdots \subseteq\left[m_{n}, M_{n}\right]$. Since $a_{k} \in\left[m_{k}, M_{k}\right]$, the claim follows.

Claim 4. For every integer $n \geqslant 2$, we have $M_{n}=S(n, n-1) /(n-1)$ and $m_{n}=S(n, n) / n$.
Proof. We use induction on $n$. The base case $n=2$ is routine. To perform the induction step, we need to prove the inequalities

$$
\begin{equation*}
\frac{S(n, n)}{n} \leqslant \frac{S(n, k)}{k} \quad \text { and } \quad \frac{S(n, k)}{k} \leqslant \frac{S(n, n-1)}{n-1} \tag{1}
\end{equation*}
$$

for every positive integer $k \leqslant n$. Clearly, these inequalities hold for $k=n$ and $k=n-1$, as $S(n, n)=S(n, n-1)>0$. In the sequel, we assume that $k<n-1$.

Now the first inequality in (1) rewrites as $n S(n, k) \geqslant k S(n, n)=k(S(n, k)+S(n-k, n-k))$, or, cancelling the terms occurring on both parts, as

$$
(n-k) S(n, k) \geqslant k S(n-k, n-k) \Longleftrightarrow S(n, k) \geqslant k \cdot \frac{S(n-k, n-k)}{n-k} .
$$

By the induction hypothesis, we have $S(n-k, n-k) /(n-k)=m_{n-k}$. By Claim 3, we get $a_{n-i} \geqslant m_{n-k}$ for all $i=1,2, \ldots, k$. Summing these $k$ inequalities we obtain

$$
S(n, k) \geqslant k m_{n-k}=k \cdot \frac{S(n-k, n-k)}{n-k},
$$

as required.
The second inequality in (1) is proved similarly. Indeed, this inequality is equivalent to

$$
\begin{aligned}
(n-1) S(n, k) \leqslant k S(n, n-1) & \Longleftrightarrow(n-k-1) S(n, k) \leqslant k S(n-k, n-k-1) \\
& \Longleftrightarrow S(n, k) \leqslant k \cdot \frac{S(n-k, n-k-1)}{n-k-1}=k M_{n-k} ;
\end{aligned}
$$

the last inequality follows again from Claim 3, as each term in $S(n, k)$ is at most $M_{n-k}$.
Now we can prove the required estimate for $a_{2018}-a_{2017}$. Set $N=2017$. By Claim 4,

$$
\begin{aligned}
a_{N+1}-a_{N} \leqslant M_{N+1}-a_{N}=\frac{S(N+1, N)}{N}-a_{N} & =\frac{a_{N}+S(N, N-1)}{N}-a_{N} \\
& =\frac{S(N, N-1)}{N}-\frac{N-1}{N} \cdot a_{N} .
\end{aligned}
$$

On the other hand, the same Claim yields

$$
a_{N} \geqslant m_{N}=\frac{S(N, N)}{N}=\frac{S(N, N-1)}{N} .
$$

Noticing that each term in $S(N, N-1)$ is at most 1 , so $S(N, N-1) \leqslant N-1$, we finally obtain

$$
a_{N+1}-a_{N} \leqslant \frac{S(N, N-1)}{N}-\frac{N-1}{N} \cdot \frac{S(N, N-1)}{N}=\frac{S(N, N-1)}{N^{2}} \leqslant \frac{N-1}{N^{2}} .
$$

Comment 1. Claim 1 in Solution 1 can be deduced from Claims 2 and 4 in Solution 2.
By Claim 4 we have $M_{n}=\frac{S(n, n-1)}{n-1}$ and $m_{n}=\frac{S(n, n)}{n}=\frac{S(n, n-1)}{n}$. It follows that $\Delta_{n}=M_{n}-m_{n}=$ $\frac{S(n, n-1)}{(n-1) n}$ and so $M_{n}=n \Delta_{n}$ and $m_{n}=(n-1) \Delta_{n}$

Similarly, $M_{n-1}=(n-1) \Delta_{n-1}$ and $m_{n-1}=(n-2) \Delta_{n-1}$. Then the inequalities $m_{n-1} \leqslant m_{n}$ and $M_{n} \leqslant M_{n-1}$ from Claim 2 write as $(n-2) \Delta_{n-1} \leqslant(n-1) \Delta_{n}$ and $n \Delta_{n} \leqslant(n-1) \Delta_{n-1}$. Hence we have the double inequality

$$
\frac{n-2}{n-1} \Delta_{n-1} \leqslant \Delta_{n} \leqslant \frac{n-1}{n} \Delta_{n-1} .
$$

Comment 2. Both solutions above discuss the properties of an arbitrary sequence satisfying the problem conditions. Instead, one may investigate only an optimal sequence which maximises the value of $a_{2018}-a_{2017}$. Here we present an observation which allows to simplify such investigation - for instance, the proofs of Claim 1 in Solution 1 and Claim 4 in Solution 2.

The sequence $\left(a_{n}\right)$ is uniquely determined by choosing, for every $n \geqslant 2$, a positive integer $k(n) \leqslant n$ such that $a_{n}=S(n, k(n)) / k(n)$. Take an arbitrary $2 \leqslant n_{0} \leqslant 2018$, and assume that all such integers $k(n)$, for $n \neq n_{0}$, are fixed. Then, for every $n$, the value of $a_{n}$ is a linear function in $a_{n_{0}}$ (whose possible values constitute some discrete subset of $\left[m_{n_{0}}, M_{n_{0}}\right]$ containing both endpoints). Hence, $a_{2018}-a_{2017}$ is also a linear function in $a_{n 0}$, so it attains its maximal value at one of the endpoints of the segment [ $m_{n_{0}}, M_{n_{0}}$ ].

This shows that, while dealing with an optimal sequence, we may assume $a_{n} \in\left\{m_{n}, M_{n}\right\}$ for all $2 \leqslant n \leqslant 2018$. Now one can easily see that, if $a_{n}=m_{n}$, then $m_{n+1}=m_{n}$ and $M_{n+1} \leqslant \frac{m_{n}+n M_{n}}{n+1}$; similar estimates hold in the case $a_{n}=M_{n}$. This already establishes Claim 1, and simplifies the inductive proof of Claim 4, both applied to an optimal sequence.

A5. Determine all functions $f:(0, \infty) \rightarrow \mathbb{R}$ satisfying

$$
\begin{equation*}
\left(x+\frac{1}{x}\right) f(y)=f(x y)+f\left(\frac{y}{x}\right) \tag{1}
\end{equation*}
$$

for all $x, y>0$.
(South Korea)
Answer: $f(x)=C_{1} x+\frac{C_{2}}{x}$ with arbitrary constants $C_{1}$ and $C_{2}$.
Solution 1. Fix a real number $a>1$, and take a new variable $t$. For the values $f(t), f\left(t^{2}\right)$, $f(a t)$ and $f\left(a^{2} t^{2}\right)$, the relation (1) provides a system of linear equations:

$$
\begin{array}{ll}
x=y=t: & \left(t+\frac{1}{t}\right) f(t)=f\left(t^{2}\right)+f(1) \\
x=\frac{t}{a}, y=a t: & \left(\frac{t}{a}+\frac{a}{t}\right) f(a t)=f\left(t^{2}\right)+f\left(a^{2}\right) \\
x=a^{2} t, y=t: & \left(a^{2} t+\frac{1}{a^{2} t}\right) f(t)=f\left(a^{2} t^{2}\right)+f\left(\frac{1}{a^{2}}\right) \\
x=y=a t: & \left(a t+\frac{1}{a t}\right) f(a t)=f\left(a^{2} t^{2}\right)+f(1) \tag{2~d}
\end{array}
$$

In order to eliminate $f\left(t^{2}\right)$, take the difference of (2a) and (2b); from (2c) and (2d) eliminate $f\left(a^{2} t^{2}\right)$; then by taking a linear combination, eliminate $f(a t)$ as well:

$$
\begin{gathered}
\left(t+\frac{1}{t}\right) f(t)-\left(\frac{t}{a}+\frac{a}{t}\right) f(a t)=f(1)-f\left(a^{2}\right) \text { and } \\
\left(a^{2} t+\frac{1}{a^{2} t}\right) f(t)-\left(a t+\frac{1}{a t}\right) f(a t)=f\left(1 / a^{2}\right)-f(1), \text { so } \\
\left(\left(a t+\frac{1}{a t}\right)\left(t+\frac{1}{t}\right)-\left(\frac{t}{a}+\frac{a}{t}\right)\left(a^{2} t+\frac{1}{a^{2} t}\right)\right) f(t) \\
=\left(a t+\frac{1}{a t}\right)\left(f(1)-f\left(a^{2}\right)\right)-\left(\frac{t}{a}+\frac{a}{t}\right)\left(f\left(1 / a^{2}\right)-f(1)\right)
\end{gathered}
$$

Notice that on the left-hand side, the coefficient of $f(t)$ is nonzero and does not depend on $t$ :

$$
\left(a t+\frac{1}{a t}\right)\left(t+\frac{1}{t}\right)-\left(\frac{t}{a}+\frac{a}{t}\right)\left(a^{2} t+\frac{1}{a^{2} t}\right)=a+\frac{1}{a}-\left(a^{3}+\frac{1}{a^{3}}\right)<0 .
$$

After dividing by this fixed number, we get

$$
\begin{equation*}
f(t)=C_{1} t+\frac{C_{2}}{t} \tag{3}
\end{equation*}
$$

where the numbers $C_{1}$ and $C_{2}$ are expressed in terms of $a, f(1), f\left(a^{2}\right)$ and $f\left(1 / a^{2}\right)$, and they do not depend on $t$.

The functions of the form (3) satisfy the equation:

$$
\left(x+\frac{1}{x}\right) f(y)=\left(x+\frac{1}{x}\right)\left(C_{1} y+\frac{C_{2}}{y}\right)=\left(C_{1} x y+\frac{C_{2}}{x y}\right)+\left(C_{1} \frac{y}{x}+C_{2} \frac{x}{y}\right)=f(x y)+f\left(\frac{y}{x}\right) .
$$

Solution 2. We start with an observation. If we substitute $x=a \neq 1$ and $y=a^{n}$ in (1), we obtain

$$
f\left(a^{n+1}\right)-\left(a+\frac{1}{a}\right) f\left(a^{n}\right)+f\left(a^{n-1}\right)=0 .
$$

For the sequence $z_{n}=a^{n}$, this is a homogeneous linear recurrence of the second order, and its characteristic polynomial is $t^{2}-\left(a+\frac{1}{a}\right) t+1=(t-a)\left(t-\frac{1}{a}\right)$ with two distinct nonzero roots, namely $a$ and $1 / a$. As is well-known, the general solution is $z_{n}=C_{1} a^{n}+C_{2}(1 / a)^{n}$ where the index $n$ can be as well positive as negative. Of course, the numbers $C_{1}$ and $C_{2}$ may depend of the choice of $a$, so in fact we have two functions, $C_{1}$ and $C_{2}$, such that

$$
\begin{equation*}
f\left(a^{n}\right)=C_{1}(a) \cdot a^{n}+\frac{C_{2}(a)}{a^{n}} \quad \text { for every } a \neq 1 \text { and every integer } n \tag{4}
\end{equation*}
$$

The relation (4) can be easily extended to rational values of $n$, so we may conjecture that $C_{1}$ and $C_{2}$ are constants, and whence $f(t)=C_{1} t+\frac{C_{2}}{t}$. As it was seen in the previous solution, such functions indeed satisfy (1).

The equation (1) is linear in $f$; so if some functions $f_{1}$ and $f_{2}$ satisfy (1) and $c_{1}, c_{2}$ are real numbers, then $c_{1} f_{1}(x)+c_{2} f_{2}(x)$ is also a solution of (1). In order to make our formulas simpler, define

$$
f_{0}(x)=f(x)-f(1) \cdot x
$$

This function is another one satisfying (1) and the extra constraint $f_{0}(1)=0$. Repeating the same argument on linear recurrences, we can write $f_{0}(a)=K(a) a^{n}+\frac{L(a)}{a^{n}}$ with some functions $K$ and $L$. By substituting $n=0$, we can see that $K(a)+L(a)=f_{0}(1)=0$ for every $a$. Hence,

$$
f_{0}\left(a^{n}\right)=K(a)\left(a^{n}-\frac{1}{a^{n}}\right)
$$

Now take two numbers $a>b>1$ arbitrarily and substitute $x=(a / b)^{n}$ and $y=(a b)^{n}$ in (1):

$$
\begin{align*}
\left(\frac{a^{n}}{b^{n}}+\frac{b^{n}}{a^{n}}\right) f_{0}\left((a b)^{n}\right) & =f_{0}\left(a^{2 n}\right)+f_{0}\left(b^{2 n}\right), \quad \text { so } \\
\left(\frac{a^{n}}{b^{n}}+\frac{b^{n}}{a^{n}}\right) K(a b)\left((a b)^{n}-\frac{1}{(a b)^{n}}\right) & =K(a)\left(a^{2 n}-\frac{1}{a^{2 n}}\right)+K(b)\left(b^{2 n}-\frac{1}{b^{2 n}}\right), \quad \text { or equivalently } \\
K(a b)\left(a^{2 n}-\frac{1}{a^{2 n}}+b^{2 n}-\frac{1}{b^{2 n}}\right) & =K(a)\left(a^{2 n}-\frac{1}{a^{2 n}}\right)+K(b)\left(b^{2 n}-\frac{1}{b^{2 n}}\right) \tag{5}
\end{align*}
$$

By dividing (5) by $a^{2 n}$ and then taking limit with $n \rightarrow+\infty$ we get $K(a b)=K(a)$. Then (5) reduces to $K(a)=K(b)$. Hence, $K(a)=K(b)$ for all $a>b>1$.

Fix $a>1$. For every $x>0$ there is some $b$ and an integer $n$ such that $1<b<a$ and $x=b^{n}$. Then

$$
f_{0}(x)=f_{0}\left(b^{n}\right)=K(b)\left(b^{n}-\frac{1}{b^{n}}\right)=K(a)\left(x-\frac{1}{x}\right) .
$$

Hence, we have $f(x)=f_{0}(x)+f(1) x=C_{1} x+\frac{C_{2}}{x}$ with $C_{1}=K(a)+f(1)$ and $C_{2}=-K(a)$.
Comment. After establishing (5), there are several variants of finishing the solution. For example, instead of taking a limit, we can obtain a system of linear equations for $K(a), K(b)$ and $K(a b)$ by substituting two positive integers $n$ in (5), say $n=1$ and $n=2$. This approach leads to a similar ending as in the first solution.

Optionally, we define another function $f_{1}(x)=f_{0}(x)-C\left(x-\frac{1}{x}\right)$ and prescribe $K(c)=0$ for another fixed $c$. Then we can choose $a b=c$ and decrease the number of terms in (5).

A6. Let $m, n \geqslant 2$ be integers. Let $f\left(x_{1}, \ldots, x_{n}\right)$ be a polynomial with real coefficients such that

$$
f\left(x_{1}, \ldots, x_{n}\right)=\left\lfloor\frac{x_{1}+\ldots+x_{n}}{m}\right\rfloor \quad \text { for every } x_{1}, \ldots, x_{n} \in\{0,1, \ldots, m-1\}
$$

Prove that the total degree of $f$ is at least $n$.
(Brazil)
Solution. We transform the problem to a single variable question by the following
Lemma. Let $a_{1}, \ldots, a_{n}$ be nonnegative integers and let $G(x)$ be a nonzero polynomial with $\operatorname{deg} G \leqslant a_{1}+\ldots+a_{n}$. Suppose that some polynomial $F\left(x_{1}, \ldots, x_{n}\right)$ satisfies

$$
F\left(x_{1}, \ldots, x_{n}\right)=G\left(x_{1}+\ldots+x_{n}\right) \quad \text { for }\left(x_{1}, \ldots, x_{n}\right) \in\left\{0,1, \ldots, a_{1}\right\} \times \ldots \times\left\{0,1, \ldots, a_{n}\right\}
$$

Then $F$ cannot be the zero polynomial, and $\operatorname{deg} F \geqslant \operatorname{deg} G$.
For proving the lemma, we will use forward differences of polynomials. If $p(x)$ is a polynomial with a single variable, then define $(\Delta p)(x)=p(x+1)-p(x)$. It is well-known that if $p$ is a nonconstant polynomial then $\operatorname{deg} \Delta p=\operatorname{deg} p-1$.

If $p\left(x_{1}, \ldots, x_{n}\right)$ is a polynomial with $n$ variables and $1 \leqslant k \leqslant n$ then let

$$
\left(\Delta_{k} p\right)\left(x_{1}, \ldots, x_{n}\right)=p\left(x_{1}, \ldots, x_{k-1}, x_{k}+1, x_{k+1}, \ldots, x_{n}\right)-p\left(x_{1}, \ldots, x_{n}\right)
$$

It is also well-known that either $\Delta_{k} p$ is the zero polynomial or $\operatorname{deg}\left(\Delta_{k} p\right) \leqslant \operatorname{deg} p-1$.
Proof of the lemma. We apply induction on the degree of $G$. If $G$ is a constant polynomial then we have $F(0, \ldots, 0)=G(0) \neq 0$, so $F$ cannot be the zero polynomial.

Suppose that $\operatorname{deg} G \geqslant 1$ and the lemma holds true for lower degrees. Since $a_{1}+\ldots+a_{n} \geqslant$ $\operatorname{deg} G>0$, at least one of $a_{1}, \ldots, a_{n}$ is positive; without loss of generality suppose $a_{1} \geqslant 1$.

Consider the polynomials $F_{1}=\Delta_{1} F$ and $G_{1}=\Delta G$. On the grid $\left\{0, \ldots, a_{1}-1\right\} \times\left\{0, \ldots, a_{2}\right\} \times$ $\ldots \times\left\{0, \ldots, a_{n}\right\}$ we have

$$
\begin{aligned}
F_{1}\left(x_{1}, \ldots, x_{n}\right) & =F\left(x_{1}+1, x_{2}, \ldots, x_{n}\right)-F\left(x_{1}, x_{2}, \ldots, x_{n}\right)= \\
& =G\left(x_{1}+\ldots+x_{n}+1\right)-G\left(x_{1}+\ldots+x_{n}\right)=G_{1}\left(x_{1}+\ldots+x_{n}\right) .
\end{aligned}
$$

Since $G$ is nonconstant, we have $\operatorname{deg} G_{1}=\operatorname{deg} G-1 \leqslant\left(a_{1}-1\right)+a_{2}+\ldots+a_{n}$. Therefore we can apply the induction hypothesis to $F_{1}$ and $G_{1}$ and conclude that $F_{1}$ is not the zero polynomial and $\operatorname{deg} F_{1} \geqslant \operatorname{deg} G_{1}$. Hence, $\operatorname{deg} F \geqslant \operatorname{deg} F_{1}+1 \geqslant \operatorname{deg} G_{1}+1=\operatorname{deg} G$. That finishes the proof.

To prove the problem statement, take the unique polynomial $g(x)$ so that $g(x)=\left\lfloor\frac{x}{m}\right\rfloor$ for $x \in\{0,1, \ldots, n(m-1)\}$ and $\operatorname{deg} g \leqslant n(m-1)$. Notice that precisely $n(m-1)+1$ values of $g$ are prescribed, so $g(x)$ indeed exists and is unique. Notice further that the constraints $g(0)=g(1)=0$ and $g(m)=1$ together enforce $\operatorname{deg} g \geqslant 2$.

By applying the lemma to $a_{1}=\ldots=a_{n}=m-1$ and the polynomials $f$ and $g$, we achieve $\operatorname{deg} f \geqslant \operatorname{deg} g$. Hence we just need a suitable lower bound on $\operatorname{deg} g$.

Consider the polynomial $h(x)=g(x+m)-g(x)-1$. The degree of $g(x+m)-g(x)$ is $\operatorname{deg} g-1 \geqslant 1$, so $\operatorname{deg} h=\operatorname{deg} g-1 \geqslant 1$, and therefore $h$ cannot be the zero polynomial. On the other hand, $h$ vanishes at the points $0,1, \ldots, n(m-1)-m$, so $h$ has at least $(n-1)(m-1)$ roots. Hence,

$$
\operatorname{deg} f \geqslant \operatorname{deg} g=\operatorname{deg} h+1 \geqslant(n-1)(m-1)+1 \geqslant n
$$

Comment 1. In the lemma we have equality for the choice $F\left(x_{1}, \ldots, x_{n}\right)=G\left(x_{1}+\ldots+x_{n}\right)$, so it indeed transforms the problem to an equivalent single-variable question.

Comment 2. If $m \geqslant 3$, the polynomial $h(x)$ can be replaced by $\Delta g$. Notice that

$$
(\Delta g)(x)=\left\{\begin{array}{ll}
1 & \text { if } x \equiv-1 \\
0 & \text { otherwise }
\end{array} \quad(\bmod m) \quad \text { for } x=0,1, \ldots, n(m-1)-1 .\right.
$$

Hence, $\Delta g$ vanishes at all integers $x$ with $0 \leqslant x<n(m-1)$ and $x \not \equiv-1(\bmod m)$. This leads to $\operatorname{deg} g \geqslant \frac{(m-1)^{2} n}{m}+1$.

If $m$ is even then this lower bound can be improved to $n(m-1)$. For $0 \leqslant N<n(m-1)$, the $(N+1)^{\text {st }}$ forward difference at $x=0$ is

$$
\begin{equation*}
\left(\Delta^{N+1}\right) g(0)=\sum_{k=0}^{N}(-1)^{N-k}\binom{N}{k}(\Delta g)(k)=\sum_{\substack{0 \leqslant k \leqslant N \\ k \equiv-1(\bmod m)}}(-1)^{N-k}\binom{N}{k} . \tag{*}
\end{equation*}
$$

Since $m$ is even, all signs in the last sum are equal; with $N=n(m-1)-1$ this proves $\Delta^{n(m-1)} g(0) \neq 0$, indicating that $\operatorname{deg} g \geqslant n(m-1)$.

However, there are infinitely many cases when all terms in (*) cancel out, for example if $m$ is an odd divisor of $n+1$. In such cases, $\operatorname{deg} f$ can be less than $n(m-1)$.

Comment 3. The lemma is closely related to the so-called
Alon-Füredi bound. Let $S_{1}, \ldots, S_{n}$ be nonempty finite sets in a field and suppose that the polynomial $P\left(x_{1}, \ldots, x_{n}\right)$ vanishes at the points of the grid $S_{1} \times \ldots \times S_{n}$, except for a single point. Then $\operatorname{deg} P \geqslant \sum_{i=1}^{n}\left(\left|S_{i}\right|-1\right)$.
(A well-known application of the Alon-Füredi bound was the former IMO problem 2007/6. Since then, this result became popular among the students and is part of the IMO training for many IMO teams.)

The proof of the lemma can be replaced by an application of the Alon-Füredi bound as follows. Let $d=\operatorname{deg} G$, and let $G_{0}$ be the unique polynomial such that $G_{0}(x)=G(x)$ for $x \in\{0,1, \ldots, d-1\}$ but $\operatorname{deg} G_{0}<d$. The polynomials $G_{0}$ and $G$ are different because they have different degrees, and they attain the same values at $0,1, \ldots, d-1$; that enforces $G_{0}(d) \neq G(d)$.

Choose some nonnegative integers $b_{1}, \ldots, b_{n}$ so that $b_{1} \leqslant a_{1}, \ldots, b_{n} \leqslant a_{n}$, and $b_{1}+\ldots+b_{n}=d$, and consider the polynomial

$$
H\left(x_{1}, \ldots, x_{n}\right)=F\left(x_{1}, \ldots, x_{n}\right)-G_{0}\left(x_{1}+\ldots+x_{n}\right)
$$

on the grid $\left\{0,1, \ldots, b_{1}\right\} \times \ldots \times\left\{0,1, \ldots, b_{n}\right\}$.
At the point $\left(b_{1}, \ldots, b_{n}\right)$ we have $H\left(b_{1}, \ldots, b_{n}\right)=G(d)-G_{0}(d) \neq 0$. At all other points of the grid we have $F=G$ and therefore $H=G-G_{0}=0$. So, by the Alon-Füredi bound, $\operatorname{deg} H \geqslant b_{1}+\ldots+b_{n}=d$. Since $\operatorname{deg} G_{0}<d$, this implies $\operatorname{deg} F=\operatorname{deg}\left(H+G_{0}\right)=\operatorname{deg} H \geqslant d=\operatorname{deg} G$.

A7. Find the maximal value of

$$
S=\sqrt[3]{\frac{a}{b+7}}+\sqrt[3]{\frac{b}{c+7}}+\sqrt[3]{\frac{c}{d+7}}+\sqrt[3]{\frac{d}{a+7}}
$$

where $a, b, c, d$ are nonnegative real numbers which satisfy $a+b+c+d=100$.
(Taiwan)
Answer: $\frac{8}{\sqrt[3]{7}}$, reached when $(a, b, c, d)$ is a cyclic permutation of $(1,49,1,49)$.
Solution 1. Since the value $8 / \sqrt[3]{7}$ is reached, it suffices to prove that $S \leqslant 8 / \sqrt[3]{7}$.
Assume that $x, y, z, t$ is a permutation of the variables, with $x \leqslant y \leqslant z \leqslant t$. Then, by the rearrangement inequality,

$$
S \leqslant\left(\sqrt[3]{\frac{x}{t+7}}+\sqrt[3]{\frac{t}{x+7}}\right)+\left(\sqrt[3]{\frac{y}{z+7}}+\sqrt[3]{\frac{z}{y+7}}\right)
$$

Claim. The first bracket above does not exceed $\sqrt[3]{\frac{x+t+14}{7}}$.
Proof. Since

$$
X^{3}+Y^{3}+3 X Y Z-Z^{3}=\frac{1}{2}(X+Y-Z)\left((X-Y)^{2}+(X+Z)^{2}+(Y+Z)^{2}\right)
$$

the inequality $X+Y \leqslant Z$ is equivalent (when $X, Y, Z \geqslant 0$ ) to $X^{3}+Y^{3}+3 X Y Z \leqslant Z^{3}$. Therefore, the claim is equivalent to

$$
\frac{x}{t+7}+\frac{t}{x+7}+3 \sqrt[3]{\frac{x t(x+t+14)}{7(x+7)(t+7)}} \leqslant \frac{x+t+14}{7}
$$

Notice that

$$
\begin{aligned}
& 3 \sqrt[3]{\frac{x t(x+t+14)}{7(x+7)(t+7)}}=3 \sqrt[3]{\frac{t(x+7)}{7(t+7)} \cdot \frac{x(t+7)}{7(x+7)} \cdot \frac{7(x+t+14)}{(t+7)(x+7)}} \\
& \qquad \leqslant \frac{t(x+7)}{7(t+7)}+\frac{x(t+7)}{7(x+7)}+\frac{7(x+t+14)}{(t+7)(x+7)}
\end{aligned}
$$

by the AM-GM inequality, so it suffices to prove

$$
\frac{x}{t+7}+\frac{t}{x+7}+\frac{t(x+7)}{7(t+7)}+\frac{x(t+7)}{7(x+7)}+\frac{7(x+t+14)}{(t+7)(x+7)} \leqslant \frac{x+t+14}{7} .
$$

A straightforward check verifies that the last inequality is in fact an equality.
The claim leads now to

$$
S \leqslant \sqrt[3]{\frac{x+t+14}{7}}+\sqrt[3]{\frac{y+z+14}{7}} \leqslant 2 \sqrt[3]{\frac{x+y+z+t+28}{14}}=\frac{8}{\sqrt[3]{7}}
$$

the last inequality being due to the AM-CM inequality (or to the fact that $\sqrt[3]{ }$ is concave on $[0, \infty)$ ).

Solution 2. We present a different proof for the estimate $S \leqslant 8 / \sqrt[3]{7}$.
Start by using Hölder's inequality:

$$
S^{3}=\left(\sum_{\mathrm{cyc}} \frac{\sqrt[6]{a} \cdot \sqrt[6]{a}}{\sqrt[3]{b+7}}\right)^{3} \leqslant \sum_{\mathrm{cyc}}(\sqrt[6]{a})^{3} \cdot \sum_{\mathrm{cyc}}(\sqrt[6]{a})^{3} \cdot \sum_{\mathrm{cyc}}\left(\frac{1}{\sqrt[3]{b+7}}\right)^{3}=\left(\sum_{\mathrm{cyc}} \sqrt{a}\right)^{2} \sum_{\mathrm{cyc}} \frac{1}{b+7} .
$$

Notice that

$$
\frac{(x-1)^{2}(x-7)^{2}}{x^{2}+7} \geqslant 0 \Longleftrightarrow x^{2}-16 x+71 \geqslant \frac{448}{x^{2}+7}
$$

yields

$$
\sum \frac{1}{b+7} \leqslant \frac{1}{448} \sum(b-16 \sqrt{b}+71)=\frac{1}{448}\left(384-16 \sum \sqrt{b}\right)=\frac{48-2 \sum \sqrt{b}}{56} .
$$

Finally,

$$
S^{3} \leqslant \frac{1}{56}\left(\sum \sqrt{a}\right)^{2}\left(48-2 \sum \sqrt{a}\right) \leqslant \frac{1}{56}\left(\frac{\sum \sqrt{a}+\sum \sqrt{a}+\left(48-2 \sum \sqrt{a}\right)}{3}\right)^{3}=\frac{512}{7}
$$

by the AM-GM inequality. The conclusion follows.
Comment. All the above works if we replace 7 and 100 with $k>0$ and $2\left(k^{2}+1\right)$, respectively; in this case, the answer becomes

$$
2 \sqrt[3]{\frac{(k+1)^{2}}{k}}
$$

Even further, a linear substitution allows to extend the solutions to a version with 7 and 100 being replaced with arbitrary positive real numbers $p$ and $q$ satisfying $q \geqslant 4 p$.

## Combinatorics

C1. Let $n \geqslant 3$ be an integer. Prove that there exists a set $S$ of $2 n$ positive integers satisfying the following property: For every $m=2,3, \ldots, n$ the set $S$ can be partitioned into two subsets with equal sums of elements, with one of subsets of cardinality $m$.
(Iceland)
Solution. We show that one of possible examples is the set

$$
S=\left\{1 \cdot 3^{k}, 2 \cdot 3^{k}: k=1,2, \ldots, n-1\right\} \cup\left\{1, \frac{3^{n}+9}{2}-1\right\} .
$$

It is readily verified that all the numbers listed above are distinct (notice that the last two are not divisible by 3 ).

The sum of elements in $S$ is

$$
\Sigma=1+\left(\frac{3^{n}+9}{2}-1\right)+\sum_{k=1}^{n-1}\left(1 \cdot 3^{k}+2 \cdot 3^{k}\right)=\frac{3^{n}+9}{2}+\sum_{k=1}^{n-1} 3^{k+1}=\frac{3^{n}+9}{2}+\frac{3^{n+1}-9}{2}=2 \cdot 3^{n} .
$$

Hence, in order to show that this set satisfies the problem requirements, it suffices to present, for every $m=2,3, \ldots, n$, an $m$-element subset $A_{m} \subset S$ whose sum of elements equals $3^{n}$.

Such a subset is

$$
A_{m}=\left\{2 \cdot 3^{k}: k=n-m+1, n-m+2, \ldots, n-1\right\} \cup\left\{1 \cdot 3^{n-m+1}\right\} .
$$

Clearly, $\left|A_{m}\right|=m$. The sum of elements in $A_{m}$ is

$$
3^{n-m+1}+\sum_{k=n-m+1}^{n-1} 2 \cdot 3^{k}=3^{n-m+1}+\frac{2 \cdot 3^{n}-2 \cdot 3^{n-m+1}}{2}=3^{n},
$$

as required.

Comment. Let us present a more general construction. Let $s_{1}, s_{2}, \ldots, s_{2 n-1}$ be a sequence of pairwise distinct positive integers satisfying $s_{2 i+1}=s_{2 i}+s_{2 i-1}$ for all $i=2,3, \ldots, n-1$. Set $s_{2 n}=s_{1}+s_{2}+$ $\cdots+s_{2 n-4}$.

Assume that $s_{2 n}$ is distinct from the other terms of the sequence. Then the set $S=\left\{s_{1}, s_{2}, \ldots, s_{2 n}\right\}$ satisfies the problem requirements. Indeed, the sum of its elements is

$$
\Sigma=\sum_{i=1}^{2 n-4} s_{i}+\left(s_{2 n-3}+s_{2 n-2}\right)+s_{2 n-1}+s_{2 n}=s_{2 n}+s_{2 n-1}+s_{2 n-1}+s_{2 n}=2 s_{2 n}+2 s_{2 n-1} .
$$

Therefore, we have

$$
\frac{\Sigma}{2}=s_{2 n}+s_{2 n-1}=s_{2 n}+s_{2 n-2}+s_{2 n-3}=s_{2 n}+s_{2 n-2}+s_{2 n-4}+s_{2 n-5}=\ldots,
$$

which shows that the required sets $A_{m}$ can be chosen as

$$
A_{m}=\left\{s_{2 n}, s_{2 n-2}, \ldots, s_{2 n-2 m+4}, s_{2 n-2 m+3}\right\} .
$$

So, the only condition to be satisfied is $s_{2 n} \notin\left\{s_{1}, s_{2}, \ldots, s_{2 n-1}\right\}$, which can be achieved in many different ways (e.g., by choosing properly the number $s_{1}$ after specifying $s_{2}, s_{3}, \ldots, s_{2 n-1}$ ).

The solution above is an instance of this general construction. Another instance, for $n>3$, is the set

$$
\left\{F_{1}, F_{2}, \ldots, F_{2 n-1}, F_{1}+\cdots+F_{2 n-4}\right\},
$$

where $F_{1}=1, F_{2}=2, F_{n+1}=F_{n}+F_{n-1}$ is the usual Fibonacci sequence.

C2. Queenie and Horst play a game on a $20 \times 20$ chessboard. In the beginning the board is empty. In every turn, Horst places a black knight on an empty square in such a way that his new knight does not attack any previous knights. Then Queenie places a white queen on an empty square. The game gets finished when somebody cannot move.

Find the maximal positive $K$ such that, regardless of the strategy of Queenie, Horst can put at least $K$ knights on the board.
(Armenia)
Answer: $K=20^{2} / 4=100$. In case of a $4 N \times 4 M$ board, the answer is $K=4 N M$.
Solution. We show two strategies, one for Horst to place at least 100 knights, and another strategy for Queenie that prevents Horst from putting more than 100 knights on the board.

A strategy for Horst: Put knights only on black squares, until all black squares get occupied.

Colour the squares of the board black and white in the usual way, such that the white and black squares alternate, and let Horst put his knights on black squares as long as it is possible. Two knights on squares of the same colour never attack each other. The number of black squares is $20^{2} / 2=200$. The two players occupy the squares in turn, so Horst will surely find empty black squares in his first 100 steps.

A strategy for Queenie: Group the squares into cycles of length 4, and after each step of Horst, occupy the opposite square in the same cycle.

Consider the squares of the board as vertices of a graph; let two squares be connected if two knights on those squares would attack each other. Notice that in a $4 \times 4$ board the squares can be grouped into 4 cycles of length 4, as shown in Figure 1. Divide the board into parts of size $4 \times 4$, and perform the same grouping in every part; this way we arrange the 400 squares of the board into 100 cycles (Figure 2).


Figure 1


Figure 2


Figure 3

The strategy of Queenie can be as follows: Whenever Horst puts a new knight to a certain square $A$, which is part of some cycle $A-B-C-D-A$, let Queenie put her queen on the opposite square $C$ in that cycle (Figure 3). From this point, Horst cannot put any knight on $A$ or $C$ because those squares are already occupied, neither on $B$ or $D$ because those squares are attacked by the knight standing on $A$. Hence, Horst can put at most one knight on each cycle, that is at most 100 knights in total.

Comment 1. Queenie's strategy can be prescribed by a simple rule: divide the board into $4 \times 4$ parts; whenever Horst puts a knight in a part $P$, Queenie reflects that square about the centre of $P$ and puts her queen on the reflected square.

Comment 2. The result remains the same if Queenie moves first. In the first turn, she may put her first queen arbitrarily. Later, if she has to put her next queen on a square that already contains a queen, she may move arbitrarily again.

C3. Let $n$ be a given positive integer. Sisyphus performs a sequence of turns on a board consisting of $n+1$ squares in a row, numbered 0 to $n$ from left to right. Initially, $n$ stones are put into square 0 , and the other squares are empty. At every turn, Sisyphus chooses any nonempty square, say with $k$ stones, takes one of those stones and moves it to the right by at most $k$ squares (the stone should stay within the board). Sisyphus' aim is to move all $n$ stones to square $n$.

Prove that Sisyphus cannot reach the aim in less than

$$
\left\lceil\frac{n}{1}\right\rceil+\left\lceil\frac{n}{2}\right\rceil+\left\lceil\frac{n}{3}\right\rceil+\cdots+\left\lceil\frac{n}{n}\right\rceil
$$

turns. (As usual, $\lceil x\rceil$ stands for the least integer not smaller than $x$.)

## (Netherlands)

Solution. The stones are indistinguishable, and all have the same origin and the same final position. So, at any turn we can prescribe which stone from the chosen square to move. We do it in the following manner. Number the stones from 1 to $n$. At any turn, after choosing a square, Sisyphus moves the stone with the largest number from this square.

This way, when stone $k$ is moved from some square, that square contains not more than $k$ stones (since all their numbers are at most $k$ ). Therefore, stone $k$ is moved by at most $k$ squares at each turn. Since the total shift of the stone is exactly $n$, at least $\lceil n / k\rceil$ moves of stone $k$ should have been made, for every $k=1,2, \ldots, n$.

By summing up over all $k=1,2, \ldots, n$, we get the required estimate.
Comment. The original submission contained the second part, asking for which values of $n$ the equality can be achieved. The answer is $n=1,2,3,4,5,7$. The Problem Selection Committee considered this part to be less suitable for the competition, due to technicalities.

C4. An anti-Pascal pyramid is a finite set of numbers, placed in a triangle-shaped array so that the first row of the array contains one number, the second row contains two numbers, the third row contains three numbers and so on; and, except for the numbers in the bottom row, each number equals the absolute value of the difference of the two numbers below it. For instance, the triangle below is an anti-Pascal pyramid with four rows, in which every integer from 1 to $1+2+3+4=10$ occurs exactly once:

\[

\]

Is it possible to form an anti-Pascal pyramid with 2018 rows, using every integer from 1 to $1+2+\cdots+2018$ exactly once?

Answer: No, it is not possible.
Solution. Let $T$ be an anti-Pascal pyramid with $n$ rows, containing every integer from 1 to $1+2+\cdots+n$, and let $a_{1}$ be the topmost number in $T$ (Figure 1). The two numbers below $a_{1}$ are some $a_{2}$ and $b_{2}=a_{1}+a_{2}$, the two numbers below $b_{2}$ are some $a_{3}$ and $b_{3}=a_{1}+a_{2}+a_{3}$, and so on and so forth all the way down to the bottom row, where some $a_{n}$ and $b_{n}=a_{1}+a_{2}+\cdots+a_{n}$ are the two neighbours below $b_{n-1}=a_{1}+a_{2}+\cdots+a_{n-1}$. Since the $a_{k}$ are $n$ pairwise distinct positive integers whose sum does not exceed the largest number in $T$, which is $1+2+\cdots+n$, it follows that they form a permutation of $1,2, \ldots, n$.


Figure 1


Figure 2

Consider now (Figure 2) the two 'equilateral' subtriangles of $T$ whose bottom rows contain the numbers to the left, respectively right, of the pair $a_{n}, b_{n}$. (One of these subtriangles may very well be empty.) At least one of these subtriangles, say $T^{\prime}$, has side length $\ell \geqslant\lceil(n-2) / 2\rceil$. Since $T^{\prime}$ obeys the anti-Pascal rule, it contains $\ell$ pairwise distinct positive integers $a_{1}^{\prime}, a_{2}^{\prime}, \ldots, a_{\ell}^{\prime}$, where $a_{1}^{\prime}$ is at the apex, and $a_{k}^{\prime}$ and $b_{k}^{\prime}=a_{1}^{\prime}+a_{2}^{\prime}+\cdots+a_{k}^{\prime}$ are the two neighbours below $b_{k-1}^{\prime}$ for each $k=2,3 \ldots, \ell$. Since the $a_{k}$ all lie outside $T^{\prime}$, and they form a permutation of $1,2, \ldots, n$, the $a_{k}^{\prime}$ are all greater than $n$. Consequently,

$$
\begin{array}{r}
b_{\ell}^{\prime} \geqslant(n+1)+(n+2)+\cdots+(n+\ell)=\frac{\ell(2 n+\ell+1)}{2} \\
\geqslant \frac{1}{2} \cdot \frac{n-2}{2}\left(2 n+\frac{n-2}{2}+1\right)=\frac{5 n(n-2)}{8},
\end{array}
$$

which is greater than $1+2+\cdots+n=n(n+1) / 2$ for $n=2018$. A contradiction.
Comment. The above estimate may be slightly improved by noticing that $b_{\ell}^{\prime} \neq b_{n}$. This implies $n(n+1) / 2=b_{n}>b_{\ell}^{\prime} \geqslant\lceil(n-2) / 2\rceil(2 n+\lceil(n-2) / 2\rceil+1) / 2$, so $n \leqslant 7$ if $n$ is odd, and $n \leqslant 12$ if $n$ is even. It seems that the largest anti-Pascal pyramid whose entries are a permutation of the integers from 1 to $1+2+\cdots+n$ has 5 rows.

C5. Let $k$ be a positive integer. The organising committee of a tennis tournament is to schedule the matches for $2 k$ players so that every two players play once, each day exactly one match is played, and each player arrives to the tournament site the day of his first match, and departs the day of his last match. For every day a player is present on the tournament, the committee has to pay 1 coin to the hotel. The organisers want to design the schedule so as to minimise the total cost of all players' stays. Determine this minimum cost.
(Russia)
Answer: The required minimum is $k\left(4 k^{2}+k-1\right) / 2$.
Solution 1. Enumerate the days of the tournament $1,2, \ldots,\binom{2 k}{2}$. Let $b_{1} \leqslant b_{2} \leqslant \cdots \leqslant b_{2 k}$ be the days the players arrive to the tournament, arranged in nondecreasing order; similarly, let $e_{1} \geqslant \cdots \geqslant e_{2 k}$ be the days they depart arranged in nonincreasing order (it may happen that a player arrives on day $b_{i}$ and departs on day $e_{j}$, where $i \neq j$ ). If a player arrives on day $b$ and departs on day $e$, then his stay cost is $e-b+1$. Therefore, the total stay cost is

$$
\Sigma=\sum_{i=1}^{2 k} e_{i}-\sum_{i=1}^{2 k} b_{i}+n=\sum_{i=1}^{2 k}\left(e_{i}-b_{i}+1\right)
$$

Bounding the total cost from below. To this end, estimate $e_{i+1}-b_{i+1}+1$. Before day $b_{i+1}$, only $i$ players were present, so at most $\binom{i}{2}$ matches could be played. Therefore, $b_{i+1} \leqslant\binom{ i}{2}+1$. Similarly, at most $\binom{i}{2}$ matches could be played after day $e_{i+1}$, so $e_{i} \geqslant\binom{ 2 k}{2}-\binom{i}{2}$. Thus,

$$
e_{i+1}-b_{i+1}+1 \geqslant\binom{ 2 k}{2}-2\binom{i}{2}=k(2 k-1)-i(i-1)
$$

This lower bound can be improved for $i>k$ : List the $i$ players who arrived first, and the $i$ players who departed last; at least $2 i-2 k$ players appear in both lists. The matches between these players were counted twice, though the players in each pair have played only once. Therefore, if $i>k$, then

$$
e_{i+1}-b_{i+1}+1 \geqslant\binom{ 2 k}{2}-2\binom{i}{2}+\binom{2 i-2 k}{2}=(2 k-i)^{2}
$$

An optimal tournament, We now describe a schedule in which the lower bounds above are all achieved simultaneously. Split players into two groups $X$ and $Y$, each of cardinality $k$. Next, partition the schedule into three parts. During the first part, the players from $X$ arrive one by one, and each newly arrived player immediately plays with everyone already present. During the third part (after all players from $X$ have already departed) the players from $Y$ depart one by one, each playing with everyone still present just before departing.

In the middle part, everyone from $X$ should play with everyone from $Y$. Let $S_{1}, S_{2}, \ldots, S_{k}$ be the players in $X$, and let $T_{1}, T_{2}, \ldots, T_{k}$ be the players in $Y$. Let $T_{1}, T_{2}, \ldots, T_{k}$ arrive in this order; after $T_{j}$ arrives, he immediately plays with all the $S_{i}, i>j$. Afterwards, players $S_{k}$, $S_{k-1}, \ldots, S_{1}$ depart in this order; each $S_{i}$ plays with all the $T_{j}, i \leqslant j$, just before his departure, and $S_{k}$ departs the day $T_{k}$ arrives. For $0 \leqslant s \leqslant k-1$, the number of matches played between $T_{k-s}$ 's arrival and $S_{k-s}$ 's departure is

$$
\sum_{j=k-s}^{k-1}(k-j)+1+\sum_{j=k-s}^{k-1}(k-j+1)=\frac{1}{2} s(s+1)+1+\frac{1}{2} s(s+3)=(s+1)^{2}
$$

Thus, if $i>k$, then the number of matches that have been played between $T_{i-k+1}$ 's arrival, which is $b_{i+1}$, and $S_{i-k+1}$ 's departure, which is $e_{i+1}$, is $(2 k-i)^{2}$; that is, $e_{i+1}-b_{i+1}+1=(2 k-i)^{2}$, showing the second lower bound achieved for all $i>k$.

If $i \leqslant k$, then the matches between the $i$ players present before $b_{i+1}$ all fall in the first part of the schedule, so there are $\binom{i}{2}$ such, and $b_{i+1}=\binom{i}{2}+1$. Similarly, after $e_{i+1}$, there are $i$ players left, all $\binom{i}{2}$ matches now fall in the third part of the schedule, and $e_{i+1}=\binom{2 k}{2}-\binom{i}{2}$. The first lower bound is therefore also achieved for all $i \leqslant k$.

Consequently, all lower bounds are achieved simultaneously, and the schedule is indeed optimal.
Evaluation. Finally, evaluate the total cost for the optimal schedule:

$$
\begin{aligned}
\Sigma & =\sum_{i=0}^{k}(k(2 k-1)-i(i-1))+\sum_{i=k+1}^{2 k-1}(2 k-i)^{2}=(k+1) k(2 k-1)-\sum_{i=0}^{k} i(i-1)+\sum_{j=1}^{k-1} j^{2} \\
& =k(k+1)(2 k-1)-k^{2}+\frac{1}{2} k(k+1)=\frac{1}{2} k\left(4 k^{2}+k-1\right) .
\end{aligned}
$$

Solution 2. Consider any tournament schedule. Label players $P_{1}, P_{2}, \ldots, P_{2 k}$ in order of their arrival, and label them again $Q_{2 k}, Q_{2 k-1}, \ldots, Q_{1}$ in order of their departure, to define a permutation $a_{1}, a_{2}, \ldots, a_{2 k}$ of $1,2, \ldots, 2 k$ by $P_{i}=Q_{a_{i}}$.

We first describe an optimal tournament for any given permutation $a_{1}, a_{2}, \ldots, a_{2 k}$ of the indices $1,2, \ldots, 2 k$. Next, we find an optimal permutation and an optimal tournament.
Optimisation for a fixed $a_{1}, \ldots, a_{2 k}$. We say that the cost of the match between $P_{i}$ and $P_{j}$ is the number of players present at the tournament when this match is played. Clearly, the Committee pays for each day the cost of the match of that day. Hence, we are to minimise the total cost of all matches.

Notice that $Q_{2 k}$ 's departure does not precede $P_{2 k}$ 's arrival. Hence, the number of players at the tournament monotonically increases (non-strictly) until it reaches $2 k$, and then monotonically decreases (non-strictly). So, the best time to schedule the match between $P_{i}$ and $P_{j}$ is either when $P_{\max (i, j)}$ arrives, or when $Q_{\max \left(a_{i}, a_{j}\right)}$ departs, in which case the cost is $\min \left(\max (i, j), \max \left(a_{i}, a_{j}\right)\right)$.

Conversely, assuming that $i>j$, if this match is scheduled between the arrivals of $P_{i}$ and $P_{i+1}$, then its cost will be exactly $i=\max (i, j)$. Similarly, one can make it cost $\max \left(a_{i}, a_{j}\right)$. Obviously, these conditions can all be simultaneously satisfied, so the minimal cost for a fixed sequence $a_{1}, a_{2}, \ldots, a_{2 k}$ is

$$
\begin{equation*}
\Sigma\left(a_{1}, \ldots, a_{2 k}\right)=\sum_{1 \leqslant i<j \leqslant 2 k} \min \left(\max (i, j), \max \left(a_{i}, a_{j}\right)\right) \tag{1}
\end{equation*}
$$

Optimising the sequence $\left(a_{i}\right)$. Optimisation hinges on the lemma below.
Lemma. If $a \leqslant b$ and $c \leqslant d$, then

$$
\begin{aligned}
\min (\max (a, x), \max (c, y))+\min & (\max (b, x), \max (d, y)) \\
\geqslant & \min (\max (a, x), \max (d, y))+\min (\max (b, x), \max (c, y))
\end{aligned}
$$

Proof. Write $a^{\prime}=\max (a, x) \leqslant \max (b, x)=b^{\prime}$ and $c^{\prime}=\max (c, y) \leqslant \max (d, y)=d^{\prime}$ and check that $\min \left(a^{\prime}, c^{\prime}\right)+\min \left(b^{\prime}, d^{\prime}\right) \geqslant \min \left(a^{\prime}, d^{\prime}\right)+\min \left(b^{\prime}, c^{\prime}\right)$.

Consider a permutation $a_{1}, a_{2}, \ldots, a_{2 k}$ such that $a_{i}<a_{j}$ for some $i<j$. Swapping $a_{i}$ and $a_{j}$ does not change the ( $i, j$ )th summand in (1), and for $\ell \notin\{i, j\}$ the sum of the $(i, \ell)$ th and the $(j, \ell)$ th summands does not increase by the Lemma. Hence the optimal value does not increase, but the number of disorders in the permutation increases. This process stops when $a_{i}=2 k+1-i$ for all $i$, so the required minimum is

$$
\begin{aligned}
S(2 k, 2 k-1, \ldots, 1) & =\sum_{1 \leqslant i<j \leqslant 2 k} \min (\max (i, j), \max (2 k+1-i, 2 k+1-j)) \\
& =\sum_{1 \leqslant i<j \leqslant 2 k} \min (j, 2 k+1-i) .
\end{aligned}
$$

The latter sum is fairly tractable and yields the stated result; we omit the details.
Comment. If the number of players is odd, say, $2 k-1$, the required minimum is $k(k-1)(4 k-1) / 2$. In this case, $|X|=k,|Y|=k-1$, the argument goes along the same lines, but some additional technicalities are to be taken care of.

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C6. Let $a$ and $b$ be distinct positive integers. The following infinite process takes place on an initially empty board.
(i) If there is at least a pair of equal numbers on the board, we choose such a pair and increase one of its components by $a$ and the other by $b$.
(ii) If no such pair exists, we write down two times the number 0 .

Prove that, no matter how we make the choices in (i), operation (ii) will be performed only finitely many times.
(Serbia)
Solution 1. We may assume $\operatorname{gcd}(a, b)=1$; otherwise we work in the same way with multiples of $d=\operatorname{gcd}(a, b)$.

Suppose that after $N$ moves of type (ii) and some moves of type (i) we have to add two new zeros. For each integer $k$, denote by $f(k)$ the number of times that the number $k$ appeared on the board up to this moment. Then $f(0)=2 N$ and $f(k)=0$ for $k<0$. Since the board contains at most one $k-a$, every second occurrence of $k-a$ on the board produced, at some moment, an occurrence of $k$; the same stands for $k-b$. Therefore,

$$
\begin{equation*}
f(k)=\left\lfloor\frac{f(k-a)}{2}\right\rfloor+\left\lfloor\frac{f(k-b)}{2}\right\rfloor, \tag{1}
\end{equation*}
$$

yielding

$$
\begin{equation*}
f(k) \geqslant \frac{f(k-a)+f(k-b)}{2}-1 . \tag{2}
\end{equation*}
$$

Since $\operatorname{gcd}(a, b)=1$, every integer $x>a b-a-b$ is expressible in the form $x=s a+t b$, with integer $s, t \geqslant 0$.

We will prove by induction on $s+t$ that if $x=s a+b t$, with $s, t$ nonnegative integers, then

$$
\begin{equation*}
f(x)>\frac{f(0)}{2^{s+t}}-2 . \tag{3}
\end{equation*}
$$

The base case $s+t=0$ is trivial. Assume now that (3) is true for $s+t=v$. Then, if $s+t=v+1$ and $x=s a+t b$, at least one of the numbers $s$ and $t$ - say $s$ - is positive, hence by (2),

$$
f(x)=f(s a+t b) \geqslant \frac{f((s-1) a+t b)}{2}-1>\frac{1}{2}\left(\frac{f(0)}{2^{s+t-1}}-2\right)-1=\frac{f(0)}{2^{s+t}}-2 .
$$

Assume now that we must perform moves of type (ii) ad infinitum. Take $n=a b-a-b$ and suppose $b>a$. Since each of the numbers $n+1, n+2, \ldots, n+b$ can be expressed in the form $s a+t b$, with $0 \leqslant s \leqslant b$ and $0 \leqslant t \leqslant a$, after moves of type (ii) have been performed $2^{a+b+1}$ times and we have to add a new pair of zeros, each $f(n+k), k=1,2, \ldots, b$, is at least 2 . In this case (1) yields inductively $f(n+k) \geqslant 2$ for all $k \geqslant 1$. But this is absurd: after a finite number of moves, $f$ cannot attain nonzero values at infinitely many points.

Solution 2. We start by showing that the result of the process in the problem does not depend on the way the operations are performed. For that purpose, it is convenient to modify the process a bit.
Claim 1. Suppose that the board initially contains a finite number of nonnegative integers, and one starts performing type ( $i$ ) moves only. Assume that one had applied $k$ moves which led to a final arrangement where no more type $(i)$ moves are possible. Then, if one starts from the same initial arrangement, performing type $(i)$ moves in an arbitrary fashion, then the process will necessarily stop at the same final arrangement

Proof. Throughout this proof, all moves are supposed to be of type (i).
Induct on $k$; the base case $k=0$ is trivial, since no moves are possible. Assume now that $k \geqslant 1$. Fix some canonical process, consisting of $k$ moves $M_{1}, M_{2}, \ldots, M_{k}$, and reaching the final arrangement $A$. Consider any sample process $m_{1}, m_{2}, \ldots$ starting with the same initial arrangement and proceeding as long as possible; clearly, it contains at least one move. We need to show that this process stops at $A$.

Let move $m_{1}$ consist in replacing two copies of $x$ with $x+a$ and $x+b$. If move $M_{1}$ does the same, we may apply the induction hypothesis to the arrangement appearing after $m_{1}$. Otherwise, the canonical process should still contain at least one move consisting in replacing $(x, x) \mapsto(x+a, x+b)$, because the initial arrangement contains at least two copies of $x$, while the final one contains at most one such.

Let $M_{i}$ be the first such move. Since the copies of $x$ are indistinguishable and no other copy of $x$ disappeared before $M_{i}$ in the canonical process, the moves in this process can be permuted as $M_{i}, M_{1}, \ldots, M_{i-1}, M_{i+1}, \ldots, M_{k}$, without affecting the final arrangement. Now it suffices to perform the move $m_{1}=M_{i}$ and apply the induction hypothesis as above.
Claim 2. Consider any process starting from the empty board, which involved exactly $n$ moves of type (ii) and led to a final arrangement where all the numbers are distinct. Assume that one starts with the board containing $2 n$ zeroes (as if $n$ moves of type (ii) were made in the beginning), applying type ( $i$ ) moves in an arbitrary way. Then this process will reach the same final arrangement.
Proof. Starting with the board with $2 n$ zeros, one may indeed model the first process mentioned in the statement of the claim, omitting the type (ii) moves. This way, one reaches the same final arrangement. Now, Claim 1 yields that this final arrangement will be obtained when type ( $i$ ) moves are applied arbitrarily.

Claim 2 allows now to reformulate the problem statement as follows: There exists an integer $n$ such that, starting from $2 n$ zeroes, one may apply type ( $i$ ) moves indefinitely.

In order to prove this, we start with an obvious induction on $s+t=k \geqslant 1$ to show that if we start with $2^{s+t}$ zeros, then we can get simultaneously on the board, at some point, each of the numbers $s a+t b$, with $s+t=k$.

Suppose now that $a<b$. Then, an appropriate use of separate groups of zeros allows us to get two copies of each of the numbers $s a+t b$, with $1 \leqslant s, t \leqslant b$.

Define $N=a b-a-b$, and notice that after representing each of numbers $N+k, 1 \leqslant k \leqslant b$, in the form $s a+t b, 1 \leqslant s, t \leqslant b$ we can get, using enough zeros, the numbers $N+1, N+2, \ldots, N+a$ and the numbers $N+1, N+2, \ldots, N+b$.

From now on we can perform only moves of type $(i)$. Indeed, if $n \geqslant N$, the occurrence of the numbers $n+1, n+2, \ldots, n+a$ and $n+1, n+2, \ldots, n+b$ and the replacement $(n+1, n+1) \mapsto$ $(n+b+1, n+a+1)$ leads to the occurrence of the numbers $n+2, n+3, \ldots, n+a+1$ and $n+2, n+3, \ldots, n+b+1$.

Comment. The proofs of Claims 1 and 2 may be extended in order to show that in fact the number of moves in the canonical process is the same as in an arbitrary sample one.

C7. Consider 2018 pairwise crossing circles no three of which are concurrent. These circles subdivide the plane into regions bounded by circular edges that meet at vertices. Notice that there are an even number of vertices on each circle. Given the circle, alternately colour the vertices on that circle red and blue. In doing so for each circle, every vertex is coloured twice once for each of the two circles that cross at that point. If the two colourings agree at a vertex, then it is assigned that colour; otherwise, it becomes yellow. Show that, if some circle contains at least 2061 yellow points, then the vertices of some region are all yellow.
(India)
Solution 1. Letting $n=2018$, we will show that, if every region has at least one non-yellow vertex, then every circle contains at most $n+\lfloor\sqrt{n-2}\rfloor-2$ yellow points. In the case at hand, the latter equals $2018+44-2=2060$, contradicting the hypothesis.

Consider the natural geometric graph $G$ associated with the configuration of $n$ circles. Fix any circle $C$ in the configuration, let $k$ be the number of yellow points on $C$, and find a suitable lower bound for the total number of yellow vertices of $G$ in terms of $k$ and $n$. It turns out that $k$ is even, and $G$ has at least

$$
\begin{equation*}
k+2\binom{k / 2}{2}+2\binom{n-k / 2-1}{2}=\frac{k^{2}}{2}-(n-2) k+(n-2)(n-1) \tag{*}
\end{equation*}
$$

yellow vertices. The proof hinges on the two lemmata below.
Lemma 1. Let two circles in the configuration cross at $x$ and $y$. Then $x$ and $y$ are either both yellow or both non-yellow.
Proof. This is because the numbers of interior vertices on the four arcs $x$ and $y$ determine on the two circles have like parities.

In particular, each circle in the configuration contains an even number of yellow vertices.
Lemma 2. If $\widehat{x y}, \overline{y z}$, and $\overrightarrow{z x}$ are circular arcs of three pairwise distinct circles in the configuration, then the number of yellow vertices in the set $\{x, y, z\}$ is odd.
Proof. Let $C_{1}, C_{2}, C_{3}$ be the three circles under consideration. Assume, without loss of generality, that $C_{2}$ and $C_{3}$ cross at $x, C_{3}$ and $C_{1}$ cross at $y$, and $C_{1}$ and $C_{2}$ cross at $z$. Let $k_{1}$, $k_{2}, k_{3}$ be the numbers of interior vertices on the three circular arcs under consideration. Since each circle in the configuration, different from the $C_{i}$, crosses the cycle $\widehat{x y} \cup \widehat{y z} \cup \overline{z x}$ at an even number of points (recall that no three circles are concurrent), and self-crossings are counted twice, the sum $k_{1}+k_{2}+k_{3}$ is even.

Let $Z_{1}$ be the colour $z$ gets from $C_{1}$ and define the other colours similarly. By the preceding, the number of bichromatic pairs in the list $\left(Z_{1}, Y_{1}\right),\left(X_{2}, Z_{2}\right),\left(Y_{3}, X_{3}\right)$ is odd. Since the total number of colour changes in a cycle $Z_{1}-Y_{1}-Y_{3}-X_{3}-X_{2}-Z_{2}-Z_{1}$ is even, the number of bichromatic pairs in the list $\left(X_{2}, X_{3}\right),\left(Y_{1}, Y_{3}\right),\left(Z_{1}, Z_{2}\right)$ is odd, and the lemma follows.

We are now in a position to prove that (*) bounds the total number of yellow vertices from below. Refer to Lemma 1 to infer that the $k$ yellow vertices on $C$ pair off to form the pairs of points where $C$ is crossed by $k / 2$ circles in the configuration. By Lemma 2, these circles cross pairwise to account for another $2\binom{k / 2}{2}$ yellow vertices. Finally, the remaining $n-k / 2-1$ circles in the configuration cross $C$ at non-yellow vertices, by Lemma 1, and Lemma 2 applies again to show that these circles cross pairwise to account for yet another $2\binom{n-k / 2-1}{2}$ yellow vertices. Consequently, there are at least (*) yellow vertices.

Next, notice that $G$ is a plane graph on $n(n-1)$ degree 4 vertices, having exactly $2 n(n-1)$ edges and exactly $n(n-1)+2$ faces (regions), the outer face inclusive (by Euler's formula for planar graphs).
Lemma 3. Each face of $G$ has equally many red and blue vertices. In particular, each face has an even number of non-yellow vertices.

Proof. Trace the boundary of a face once in circular order, and consider the colours each vertex is assigned in the colouring of the two circles that cross at that vertex, to infer that colours of non-yellow vertices alternate.

Consequently, if each region has at least one non-yellow vertex, then it has at least two such. Since each vertex of $G$ has degree 4, consideration of vertex-face incidences shows that $G$ has at least $n(n-1) / 2+1$ non-yellow vertices, and hence at most $n(n-1) / 2-1$ yellow vertices. (In fact, Lemma 3 shows that there are at least $n(n-1) / 4+1 / 2$ red, respectively blue, vertices.)

Finally, recall the lower bound (*) for the total number of yellow vertices in $G$, to write $n(n-1) / 2-1 \geqslant k^{2} / 2-(n-2) k+(n-2)(n-1)$, and conclude that $k \leqslant n+\lfloor\sqrt{n-2}\rfloor-2$, as claimed in the first paragraph.

Solution 2. The first two lemmata in Solution 1 show that the circles in the configuration split into two classes: Consider any circle $C$ along with all circles that cross $C$ at yellow points to form one class; the remaining circles then form the other class. Lemma 2 shows that any pair of circles in the same class cross at yellow points; otherwise, they cross at non-yellow points.

Call the circles from the two classes white and black, respectively. Call a region yellow if its vertices are all yellow. Let $w$ and $b$ be the numbers of white and black circles, respectively; clearly, $w+b=n$. Assume that $w \geqslant b$, and that there is no yellow region. Clearly, $b \geqslant 1$, otherwise each region is yellow. The white circles subdivide the plane into $w(w-1)+2$ larger regions - call them white. The white regions (or rather their boundaries) subdivide each black circle into black arcs. Since there are no yellow regions, each white region contains at least one black arc.

Consider any white region; let it contain $t \geqslant 1$ black arcs. We claim that the number of points at which these $t$ arcs cross does not exceed $t-1$. To prove this, consider a multigraph whose vertices are these black arcs, two vertices being joined by an edge for each point at which the corresponding arcs cross. If this graph had more than $t-1$ edges, it would contain a cycle, since it has $t$ vertices; this cycle would correspond to a closed contour formed by black sub-arcs, lying inside the region under consideration. This contour would, in turn, define at least one yellow region, which is impossible.

Let $t_{i}$ be the number of black arcs inside the $i^{\text {th }}$ white region. The total number of black arcs is $\sum_{i} t_{i}=2 w b$, and they cross at $2\binom{b}{2}=b(b-1)$ points. By the preceding,

$$
b(b-1) \leqslant \sum_{i=1}^{w^{2}-w+2}\left(t_{i}-1\right)=\sum_{i=1}^{w^{2}-w+2} t_{i}-\left(w^{2}-w+2\right)=2 w b-\left(w^{2}-w+2\right)
$$

or, equivalently, $(w-b)^{2} \leqslant w+b-2=n-2$, which is the case if and only if $w-b \leqslant\lfloor\sqrt{n-2}\rfloor$. Consequently, $b \leqslant w \leqslant(n+\lfloor\sqrt{n-2}\rfloor) / 2$, so there are at most $2(w-1) \leqslant n+\lfloor\sqrt{n-2}\rfloor-2$ yellow vertices on each circle - a contradiction.

## Geometry

G1. Let $A B C$ be an acute-angled triangle with circumcircle $\Gamma$. Let $D$ and $E$ be points on the segments $A B$ and $A C$, respectively, such that $A D=A E$. The perpendicular bisectors of the segments $B D$ and $C E$ intersect the small arcs $\widehat{A B}$ and $\widehat{A C}$ at points $F$ and $G$ respectively. Prove that $D E \| F G$.
(Greece)
Solution 1. In the sequel, all the considered arcs are small arcs.
Let $P$ be the midpoint of the arc $\widehat{B C}$. Then $A P$ is the bisector of $\angle B A C$, hence, in the isosceles triangle $A D E, A P \perp D E$. So, the statement of the problem is equivalent to $A P \perp F G$.

In order to prove this, let $K$ be the second intersection of $\Gamma$ with $F D$. Then the triangle $F B D$ is isosceles, therefore

$$
\angle A K F=\angle A B F=\angle F D B=\angle A D K,
$$

yielding $A K=A D$. In the same way, denoting by $L$ the second intersection of $\Gamma$ with $G E$, we get $A L=A E$. This shows that $A K=A L$.


Now $\angle F B D=\angle F D B$ gives $\overparen{A F}=\overparen{B F}+\overparen{A K}=\overparen{B F}+\overparen{A L}$, hence $\overparen{B F}=\overparen{L F}$. In a similar way, we get $\widehat{C G}=\widehat{G K}$. This yields

$$
\angle(A P, F G)=\frac{\widehat{A F}+\widehat{P G}}{2}=\frac{\widehat{A L}+\widehat{L F}+\widehat{P C}+\widehat{C G}}{2}=\frac{\widehat{K L}+\widehat{L B}+\widehat{B C}+\widehat{C K}}{4}=90^{\circ} .
$$

Solution 2. Let $Z=A B \cap F G, T=A C \cap F G$. It suffices to prove that $\angle A T Z=\angle A Z T$.
Let $X$ be the point for which $F X A D$ is a parallelogram. Then

$$
\angle F X A=\angle F D A=180^{\circ}-\angle F D B=180^{\circ}-\angle F B D,
$$

where in the last equality we used that $F D=F B$. It follows that the quadrilateral $B F X A$ is cyclic, so $X$ lies on $\Gamma$.


Analogously, if $Y$ is the point for which $G Y A E$ is a parallelogram, then $Y$ lies on $\Gamma$. So the quadrilateral $X F G Y$ is cyclic and $F X=A D=A E=G Y$, hence $X F G Y$ is an isosceles trapezoid.

Now, by $X F \| A Z$ and $Y G \| A T$, it follows that $\angle A T Z=\angle Y G F=\angle X F G=\angle A Z T$.
Solution 3. As in the first solution, we prove that $F G \perp A P$, where $P$ is the midpoint of the small arc $\widehat{B C}$.

Let $O$ be the circumcentre of the triangle $A B C$, and let $M$ and $N$ be the midpoints of the small $\operatorname{arcs} \widehat{A B}$ and $\overparen{A C}$, respectively. Then $O M$ and $O N$ are the perpendicular bisectors of $A B$ and $A C$, respectively.


The distance $d$ between $O M$ and the perpendicular bisector of $B D$ is $\frac{1}{2} A B-\frac{1}{2} B D=\frac{1}{2} A D$, hence it is equal to the distance between $O N$ and the perpendicular bisector of $C E$.

This shows that the isosceles trapezoid determined by the diameter $\delta$ of $\Gamma$ through $M$ and the chord parallel to $\delta$ through $F$ is congruent to the isosceles trapezoid determined by the diameter $\delta^{\prime}$ of $\Gamma$ through $N$ and the chord parallel to $\delta^{\prime}$ through $G$. Therefore $M F=N G$, yielding $M N \| F G$.

Now

$$
\angle(M N, A P)=\frac{1}{2}(\widetilde{A M}+\overparen{P C}+\overparen{C N})=\frac{1}{4}(\widetilde{A B}+\overparen{B C}+\overparen{C A})=90^{\circ}
$$

hence $M N \perp A P$, and the conclusion follows.

G2. Let $A B C$ be a triangle with $A B=A C$, and let $M$ be the midpoint of $B C$. Let $P$ be a point such that $P B<P C$ and $P A$ is parallel to $B C$. Let $X$ and $Y$ be points on the lines $P B$ and $P C$, respectively, so that $B$ lies on the segment $P X, C$ lies on the segment $P Y$, and $\angle P X M=\angle P Y M$. Prove that the quadrilateral $A P X Y$ is cyclic.
(Australia)
Solution. Since $A B=A C, A M$ is the perpendicular bisector of $B C$, hence $\angle P A M=$ $\angle A M C=90^{\circ}$.


Now let $Z$ be the common point of $A M$ and the perpendicular through $Y$ to $P C$ (notice that $Z$ lies on to the ray $A M$ beyond $M$ ). We have $\angle P A Z=\angle P Y Z=90^{\circ}$. Thus the points $P, A, Y$, and $Z$ are concyclic.

Since $\angle C M Z=\angle C Y Z=90^{\circ}$, the quadrilateral $C Y Z M$ is cyclic, hence $\angle C Z M=$ $\angle C Y M$. By the condition in the statement, $\angle C Y M=\angle B X M$, and, by symmetry in $Z M$, $\angle C Z M=\angle B Z M$. Therefore, $\angle B X M=\angle B Z M$. It follows that the points $B, X, Z$, and $M$ are concyclic, hence $\angle B X Z=180^{\circ}-\angle B M Z=90^{\circ}$.

Finally, we have $\angle P X Z=\angle P Y Z=\angle P A Z=90^{\circ}$, hence the five points $P, A, X, Y, Z$ are concyclic. In particular, the quadrilateral $A P X Y$ is cyclic, as required.

Comment 1. Clearly, the key point $Z$ from the solution above can be introduced in several different ways, e.g., as the second meeting point of the circle $C M Y$ and the line $A M$, or as the second meeting point of the circles $C M Y$ and $B M X$, etc.

For some of definitions of $Z$ its location is not obvious. For instance, if $Z$ is defined as a common point of $A M$ and the perpendicular through $X$ to $P X$, it is not clear that $Z$ lies on the ray $A M$ beyond $M$. To avoid such slippery details some more restrictions on the construction may be required.

Comment 2. Let us discuss a connection to the Miquel point of a cyclic quadrilateral. Set $X^{\prime}=$ $M X \cap P C, Y^{\prime}=M Y \cap P B$, and $Q=X Y \cap X^{\prime} Y^{\prime}$ (see the figure below).

We claim that $B C \| P Q$. (One way of proving this is the following. Notice that the quadruple of lines $P X, P M, P Y, P Q$ is harmonic, hence the quadruple $B, M, C, P Q \cap B C$ of their intersection points with $B C$ is harmonic. Since $M$ is the midpoint of $B C, P Q \cap B C$ is an ideal point, i.e., $P Q \| B C$.)

It follows from the given equality $\angle P X M=\angle P Y M$ that the quadrilateral $X Y X^{\prime} Y^{\prime}$ is cyclic. Note that $A$ is the projection of $M$ onto $P Q$. By a known description, $A$ is the Miquel point for the sidelines $X Y, X Y^{\prime}, X^{\prime} Y, X^{\prime} Y^{\prime}$. In particular, the circle $P X Y$ passes through $A$.


Comment 3. An alternative approach is the following. One can note that the (oriented) lengths of the segments $C Y$ and $B X$ are both linear functions of a parameter $t=\cot \angle P X M$. As $t$ varies, the intersection point $S$ of the perpendicular bisectors of $P X$ and $P Y$ traces a fixed line, thus the family of circles $P X Y$ has a fixed common point (other than $P$ ). By checking particular cases, one can show that this fixed point is $A$.

Comment 4. The problem states that $\angle P X M=\angle P Y M$ implies that $A P X Y$ is cyclic. The original submission claims that these two conditions are in fact equivalent. The Problem Selection Committee omitted the converse part, since it follows easily from the direct one, by reversing arguments.

G3. A circle $\omega$ of radius 1 is given. A collection $T$ of triangles is called good, if the following conditions hold:
(i) each triangle from $T$ is inscribed in $\omega$;
(ii) no two triangles from $T$ have a common interior point.

Determine all positive real numbers $t$ such that, for each positive integer $n$, there exists a good collection of $n$ triangles, each of perimeter greater than $t$.
(South Africa)
Answer: $t \in(0,4]$.
Solution. First, we show how to construct a good collection of $n$ triangles, each of perimeter greater than 4 . This will show that all $t \leqslant 4$ satisfy the required conditions.

Construct inductively an $(n+2)$-gon $B A_{1} A_{2} \ldots A_{n} C$ inscribed in $\omega$ such that $B C$ is a diameter, and $B A_{1} A_{2}, B A_{2} A_{3}, \ldots, B A_{n-1} A_{n}, B A_{n} C$ is a good collection of $n$ triangles. For $n=1$, take any triangle $B A_{1} C$ inscribed in $\omega$ such that $B C$ is a diameter; its perimeter is greater than $2 B C=4$. To perform the inductive step, assume that the $(n+2)$-gon $B A_{1} A_{2} \ldots A_{n} C$ is already constructed. Since $A_{n} B+A_{n} C+B C>4$, one can choose a point $A_{n+1}$ on the small $\operatorname{arc} \widehat{C A_{n}}$, close enough to $C$, so that $A_{n} B+A_{n} A_{n+1}+B A_{n+1}$ is still greater than 4. Thus each of these new triangles $B A_{n} A_{n+1}$ and $B A_{n+1} C$ has perimeter greater than 4, which completes the induction step.


We proceed by showing that no $t>4$ satisfies the conditions of the problem. To this end, we assume that there exists a good collection $T$ of $n$ triangles, each of perimeter greater than $t$, and then bound $n$ from above.

Take $\varepsilon>0$ such that $t=4+2 \varepsilon$.
Claim. There exists a positive constant $\sigma=\sigma(\varepsilon)$ such that any triangle $\Delta$ with perimeter $2 s \geqslant 4+2 \varepsilon$, inscribed in $\omega$, has area $S(\Delta)$ at least $\sigma$.
Proof. Let $a, b, c$ be the side lengths of $\Delta$. Since $\Delta$ is inscribed in $\omega$, each side has length at most 2. Therefore, $s-a \geqslant(2+\varepsilon)-2=\varepsilon$. Similarly, $s-b \geqslant \varepsilon$ and $s-c \geqslant \varepsilon$. By Heron's formula, $S(\Delta)=\sqrt{s(s-a)(s-b)(s-c)} \geqslant \sqrt{(2+\varepsilon) \varepsilon^{3}}$. Thus we can set $\sigma(\varepsilon)=\sqrt{(2+\varepsilon) \varepsilon^{3}}$.

Now we see that the total area $S$ of all triangles from $T$ is at least $n \sigma(\varepsilon)$. On the other hand, $S$ does not exceed the area of the disk bounded by $\omega$. Thus $n \sigma(\varepsilon) \leqslant \pi$, which means that $n$ is bounded from above.

Comment 1. One may prove the Claim using the formula $S=\frac{a b c}{4 R}$ instead of Heron's formula.
Comment 2. In the statement of the problem condition $(i)$ could be replaced by a weaker one: each triangle from $T$ lies within $\omega$. This does not affect the solution above, but reduces the number of ways to prove the Claim.

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G4. A point $T$ is chosen inside a triangle $A B C$. Let $A_{1}, B_{1}$, and $C_{1}$ be the reflections of $T$ in $B C, C A$, and $A B$, respectively. Let $\Omega$ be the circumcircle of the triangle $A_{1} B_{1} C_{1}$. The lines $A_{1} T, B_{1} T$, and $C_{1} T$ meet $\Omega$ again at $A_{2}, B_{2}$, and $C_{2}$, respectively. Prove that the lines $A A_{2}, B B_{2}$, and $C C_{2}$ are concurrent on $\Omega$.
(Mongolia)
Solution. By $\Varangle(\ell, n)$ we always mean the directed angle of the lines $\ell$ and $n$, taken modulo $180^{\circ}$.
Let $C C_{2}$ meet $\Omega$ again at $K$ (as usual, if $C C_{2}$ is tangent to $\Omega$, we set $T=C_{2}$ ). We show that the line $B B_{2}$ contains $K$; similarly, $A A_{2}$ will also pass through $K$. For this purpose, it suffices to prove that

$$
\begin{equation*}
\Varangle\left(C_{2} C, C_{2} A_{1}\right)=\Varangle\left(B_{2} B, B_{2} A_{1}\right) . \tag{1}
\end{equation*}
$$

By the problem condition, $C B$ and $C A$ are the perpendicular bisectors of $T A_{1}$ and $T B_{1}$, respectively. Hence, $C$ is the circumcentre of the triangle $A_{1} T B_{1}$. Therefore,

$$
\Varangle\left(C A_{1}, C B\right)=\Varangle(C B, C T)=\Varangle\left(B_{1} A_{1}, B_{1} T\right)=\Varangle\left(B_{1} A_{1}, B_{1} B_{2}\right) .
$$

In circle $\Omega$ we have $\Varangle\left(B_{1} A_{1}, B_{1} B_{2}\right)=\Varangle\left(C_{2} A_{1}, C_{2} B_{2}\right)$. Thus,

$$
\begin{equation*}
\Varangle\left(C A_{1}, C B\right)=\Varangle\left(B_{1} A_{1}, B_{1} B_{2}\right)=\Varangle\left(C_{2} A_{1}, C_{2} B_{2}\right) . \tag{2}
\end{equation*}
$$

Similarly, we get

$$
\begin{equation*}
\Varangle\left(B A_{1}, B C\right)=\Varangle\left(C_{1} A_{1}, C_{1} C_{2}\right)=\Varangle\left(B_{2} A_{1}, B_{2} C_{2}\right) . \tag{3}
\end{equation*}
$$

The two obtained relations yield that the triangles $A_{1} B C$ and $A_{1} B_{2} C_{2}$ are similar and equioriented, hence

$$
\frac{A_{1} B_{2}}{A_{1} B}=\frac{A_{1} C_{2}}{A_{1} C} \quad \text { and } \quad \Varangle\left(A_{1} B, A_{1} C\right)=\Varangle\left(A_{1} B_{2}, A_{1} C_{2}\right) .
$$

The second equality may be rewritten as $\Varangle\left(A_{1} B, A_{1} B_{2}\right)=\Varangle\left(A_{1} C, A_{1} C_{2}\right)$, so the triangles $A_{1} B B_{2}$ and $A_{1} C C_{2}$ are also similar and equioriented. This establishes (1).


Comment 1. In fact, the triangle $A_{1} B C$ is an image of $A_{1} B_{2} C_{2}$ under a spiral similarity centred at $A_{1}$; in this case, the triangles $A B B_{2}$ and $A C C_{2}$ are also spirally similar with the same centre.

Comment 2. After obtaining (2) and (3), one can finish the solution in different ways.
For instance, introducing the point $X=B C \cap B_{2} C_{2}$, one gets from these relations that the 4 -tuples $\left(A_{1}, B, B_{2}, X\right)$ and $\left(A_{1}, C, C_{2}, X\right)$ are both cyclic. Therefore, $K$ is the Miquel point of the lines $B B_{2}$, $C C_{2}, B C$, and $B_{2} C_{2}$; this yields that the meeting point of $B B_{2}$ and $C C_{2}$ lies on $\Omega$.

Yet another way is to show that the points $A_{1}, B, C$, and $K$ are concyclic, as

$$
\Varangle\left(K C, K A_{1}\right)=\Varangle\left(B_{2} C_{2}, B_{2} A_{1}\right)=\Varangle\left(B C, B A_{1}\right) .
$$

By symmetry, the second point $K^{\prime}$ of intersection of $B B_{2}$ with $\Omega$ is also concyclic to $A_{1}, B$, and $C$, hence $K^{\prime}=K$.


Comment 3. The requirement that the common point of the lines $A A_{2}, B B_{2}$, and $C C_{2}$ should lie on $\Omega$ may seem to make the problem easier, since it suggests some approaches. On the other hand, there are also different ways of showing that the lines $A A_{2}, B B_{2}$, and $C C_{2}$ are just concurrent.

In particular, the problem conditions yield that the lines $A_{2} T, B_{2} T$, and $C_{2} T$ are perpendicular to the corresponding sides of the triangle $A B C$. One may show that the lines $A T, B T$, and $C T$ are also perpendicular to the corresponding sides of the triangle $A_{2} B_{2} C_{2}$, i.e., the triangles $A B C$ and $A_{2} B_{2} C_{2}$ are orthologic, and their orthology centres coincide. It is known that such triangles are also perspective, i.e. the lines $A A_{2}, B B_{2}$, and $C C_{2}$ are concurrent (in projective sense).

To show this mutual orthology, one may again apply angle chasing, but there are also other methods. Let $A^{\prime}, B^{\prime}$, and $C^{\prime}$ be the projections of $T$ onto the sides of the triangle $A B C$. Then $A_{2} T \cdot T A^{\prime}=$ $B_{2} T \cdot T B^{\prime}=C_{2} T \cdot T C^{\prime}$, since all three products equal (minus) half the power of $T$ with respect to $\Omega$. This means that $A_{2}, B_{2}$, and $C_{2}$ are the poles of the sidelines of the triangle $A B C$ with respect to some circle centred at $T$ and having pure imaginary radius (in other words, the reflections of $A_{2}, B_{2}$, and $C_{2}$ in $T$ are the poles of those sidelines with respect to some regular circle centred at $T$ ). Hence, dually, the vertices of the triangle $A B C$ are also the poles of the sidelines of the triangle $A_{2} B_{2} C_{2}$.

G5. Let $A B C$ be a triangle with circumcircle $\omega$ and incentre $I$. A line $\ell$ intersects the lines $A I, B I$, and $C I$ at points $D, E$, and $F$, respectively, distinct from the points $A, B, C$, and $I$. The perpendicular bisectors $x, y$, and $z$ of the segments $A D, B E$, and $C F$, respectively determine a triangle $\Theta$. Show that the circumcircle of the triangle $\Theta$ is tangent to $\omega$.
(Denmark)

Preamble. Let $X=y \cap z, Y=x \cap z, Z=x \cap y$ and let $\Omega$ denote the circumcircle of the triangle $X Y Z$. Denote by $X_{0}, Y_{0}$, and $Z_{0}$ the second intersection points of $A I, B I$ and $C I$, respectively, with $\omega$. It is known that $Y_{0} Z_{0}$ is the perpendicular bisector of $A I, Z_{0} X_{0}$ is the perpendicular bisector of $B I$, and $X_{0} Y_{0}$ is the perpendicular bisector of $C I$. In particular, the triangles $X Y Z$ and $X_{0} Y_{0} Z_{0}$ are homothetic, because their corresponding sides are parallel.

The solutions below mostly exploit the following approach. Consider the triangles $X Y Z$ and $X_{0} Y_{0} Z_{0}$, or some other pair of homothetic triangles $\Delta$ and $\delta$ inscribed into $\Omega$ and $\omega$, respectively. In order to prove that $\Omega$ and $\omega$ are tangent, it suffices to show that the centre $T$ of the homothety taking $\Delta$ to $\delta$ lies on $\omega$ (or $\Omega$ ), or, in other words, to show that $\Delta$ and $\delta$ are perspective (i.e., the lines joining corresponding vertices are concurrent), with their perspector lying on $\omega$ (or $\Omega$ ).

We use directed angles throughout all the solutions.

## Solution 1.

Claim 1. The reflections $\ell_{a}, \ell_{b}$ and $\ell_{c}$ of the line $\ell$ in the lines $x, y$, and $z$, respectively, are concurrent at a point $T$ which belongs to $\omega$.


Proof. Notice that $\Varangle\left(\ell_{b}, \ell_{c}\right)=\Varangle\left(\ell_{b}, \ell\right)+\Varangle\left(\ell, \ell_{c}\right)=2 \Varangle(y, \ell)+2 \Varangle(\ell, z)=2 \Varangle(y, z)$. But $y \perp B I$ and $z \perp C I$ implies $\Varangle(y, z)=\Varangle(B I, I C)$, so, since $2 \Varangle(B I, I C)=\Varangle(B A, A C)$, we obtain

$$
\begin{equation*}
\Varangle\left(\ell_{b}, \ell_{c}\right)=\Varangle(B A, A C) . \tag{1}
\end{equation*}
$$

Since $A$ is the reflection of $D$ in $x, A$ belongs to $\ell_{a}$; similarly, $B$ belongs to $\ell_{b}$. Then (1) shows that the common point $T^{\prime}$ of $\ell_{a}$ and $\ell_{b}$ lies on $\omega$; similarly, the common point $T^{\prime \prime}$ of $\ell_{c}$ and $\ell_{b}$ lies on $\omega$.

If $B \notin \ell_{a}$ and $B \notin \ell_{c}$, then $T^{\prime}$ and $T^{\prime \prime}$ are the second point of intersection of $\ell_{b}$ and $\omega$, hence they coincide. Otherwise, if, say, $B \in \ell_{c}$, then $\ell_{c}=B C$, so $\Varangle(B A, A C)=\Varangle\left(\ell_{b}, \ell_{c}\right)=\Varangle\left(\ell_{b}, B C\right)$, which shows that $\ell_{b}$ is tangent at $B$ to $\omega$ and $T^{\prime}=T^{\prime \prime}=B$. So $T^{\prime}$ and $T^{\prime \prime}$ coincide in all the cases, and the conclusion of the claim follows.

Now we prove that $X, X_{0}, T$ are collinear. Denote by $D_{b}$ and $D_{c}$ the reflections of the point $D$ in the lines $y$ and $z$, respectively. Then $D_{b}$ lies on $\ell_{b}, D_{c}$ lies on $\ell_{c}$, and

$$
\begin{aligned}
\Varangle\left(D_{b} X, X D_{c}\right) & =\Varangle\left(D_{b} X, D X\right)+\Varangle\left(D X, X D_{c}\right)=2 \Varangle(y, D X)+2 \Varangle(D X, z)=2 \Varangle(y, z) \\
& =\Varangle(B A, A C)=\Varangle(B T, T C),
\end{aligned}
$$

hence the quadrilateral $X D_{b} T D_{c}$ is cyclic. Notice also that since $X D_{b}=X D=X D_{c}$, the points $D, D_{b}, D_{c}$ lie on a circle with centre $X$. Using in this circle the diameter $D_{c} D_{c}^{\prime}$ yields $\Varangle\left(D_{b} D_{c}, D_{c} X\right)=90^{\circ}+\Varangle\left(D_{b} D_{c}^{\prime}, D_{c}^{\prime} X\right)=90^{\circ}+\Varangle\left(D_{b} D, D D_{c}\right)$. Therefore,

$$
\begin{gathered}
\Varangle\left(\ell_{b}, X T\right)=\Varangle\left(D_{b} T, X T\right)=\Varangle\left(D_{b} D_{c}, D_{c} X\right)=90^{\circ}+\Varangle\left(D_{b} D, D D_{c}\right) \\
=90^{\circ}+\Varangle(B I, I C)=\Varangle(B A, A I)=\Varangle\left(B A, A X_{0}\right)=\Varangle\left(B T, T X_{0}\right)=\Varangle\left(\ell_{b}, X_{0} T\right),
\end{gathered}
$$

so the points $X, X_{0}, T$ are collinear. By a similar argument, $Y, Y_{0}, T$ and $Z, Z_{0}, T$ are collinear. As mentioned in the preamble, the statement of the problem follows.

Comment 1. After proving Claim 1 one may proceed in another way. As it was shown, the reflections of $\ell$ in the sidelines of $X Y Z$ are concurrent at $T$. Thus $\ell$ is the Steiner line of $T$ with respect to $\triangle X Y Z$ (that is the line containing the reflections $T_{a}, T_{b}, T_{c}$ of $T$ in the sidelines of $X Y Z$ ). The properties of the Steiner line imply that $T$ lies on $\Omega$, and $\ell$ passes through the orthocentre $H$ of the triangle $X Y Z$.


Let $H_{a}, H_{b}$, and $H_{c}$ be the reflections of the point $H$ in the lines $x, y$, and $z$, respectively. Then the triangle $H_{a} H_{b} H_{c}$ is inscribed in $\Omega$ and homothetic to $A B C$ (by an easy angle chasing). Since $H_{a} \in \ell_{a}, H_{b} \in \ell_{b}$, and $H_{c} \in \ell_{c}$, the triangles $H_{a} H_{b} H_{c}$ and $A B C$ form a required pair of triangles $\Delta$ and $\delta$ mentioned in the preamble.

Comment 2. The following observation shows how one may guess the description of the tangency point $T$ from Solution 1.

Let us fix a direction and move the line $\ell$ parallel to this direction with constant speed.
Then the points $D, E$, and $F$ are moving with constant speeds along the lines $A I, B I$, and $C I$, respectively. In this case $x, y$, and $z$ are moving with constant speeds, defining a family of homothetic triangles $X Y Z$ with a common centre of homothety $T$. Notice that the triangle $X_{0} Y_{0} Z_{0}$ belongs to this family (for $\ell$ passing through $I$ ). We may specify the location of $T$ considering the degenerate case when $x, y$, and $z$ are concurrent. In this degenerate case all the lines $x, y, z, \ell, \ell_{a}, \ell_{b}, \ell_{c}$ have a common point. Note that the lines $\ell_{a}, \ell_{b}, \ell_{c}$ remain constant as $\ell$ is moving (keeping its direction). Thus $T$ should be the common point of $\ell_{a}, \ell_{b}$, and $\ell_{c}$, lying on $\omega$.

Solution 2. As mentioned in the preamble, it is sufficient to prove that the centre $T$ of the homothety taking $X Y Z$ to $X_{0} Y_{0} Z_{0}$ belongs to $\omega$. Thus, it suffices to prove that $\Varangle\left(T X_{0}, T Y_{0}\right)=$ $\Varangle\left(Z_{0} X_{0}, Z_{0} Y_{0}\right)$, or, equivalently, $\Varangle\left(X X_{0}, Y Y_{0}\right)=\Varangle\left(Z_{0} X_{0}, Z_{0} Y_{0}\right)$.

Recall that $Y Z$ and $Y_{0} Z_{0}$ are the perpendicular bisectors of $A D$ and $A I$, respectively. Then, the vector $\vec{x}$ perpendicular to $Y Z$ and shifting the line $Y_{0} Z_{0}$ to $Y Z$ is equal to $\frac{1}{2} \overrightarrow{I D}$. Define the shifting vectors $\vec{y}=\frac{1}{2} \overrightarrow{I E}, \vec{z}=\frac{1}{2} \overrightarrow{I F}$ similarly. Consider now the triangle $U V W$ formed by the perpendiculars to $A I, B I$, and $C I$ through $D, E$, and $F$, respectively (see figure below). This is another triangle whose sides are parallel to the corresponding sides of $X Y Z$.
Claim 2. $\overrightarrow{I U}=2 \overrightarrow{X_{0} X}, \overrightarrow{I V}=2 \overrightarrow{Y_{0} Y}, \overrightarrow{I W}=2 \overrightarrow{Z_{0} Z}$.
Proof. We prove one of the relations, the other proofs being similar. To prove the equality of two vectors it suffices to project them onto two non-parallel axes and check that their projections are equal.

The projection of $\overrightarrow{X_{0} X}$ onto $I B$ equals $\vec{y}$, while the projection of $\overrightarrow{I U}$ onto $I B$ is $\overrightarrow{I E}=2 \vec{y}$. The projections onto the other axis $I C$ are $\vec{z}$ and $\overrightarrow{I F}=2 \vec{z}$. Then $\overrightarrow{I U}=2 \overrightarrow{X_{0} X}$ follows.

Notice that the line $\ell$ is the Simson line of the point $I$ with respect to the triangle $U V W$; thus $U, V, W$, and $I$ are concyclic. It follows from Claim 2 that $\Varangle\left(X X_{0}, Y Y_{0}\right)=\Varangle(I U, I V)=$ $\Varangle(W U, W V)=\Varangle\left(Z_{0} X_{0}, Z_{0} Y_{0}\right)$, and we are done.


Solution 3. Let $I_{a}, I_{b}$, and $I_{c}$ be the excentres of triangle $A B C$ corresponding to $A, B$, and $C$, respectively. Also, let $u, v$, and $w$ be the lines through $D, E$, and $F$ which are perpendicular to $A I, B I$, and $C I$, respectively, and let $U V W$ be the triangle determined by these lines, where $u=V W, v=U W$ and $w=U V$ (see figure above).

Notice that the line $u$ is the reflection of $I_{b} I_{c}$ in the line $x$, because $u, x$, and $I_{b} I_{c}$ are perpendicular to $A D$ and $x$ is the perpendicular bisector of $A D$. Likewise, $v$ and $I_{a} I_{c}$ are reflections of each other in $y$, while $w$ and $I_{a} I_{b}$ are reflections of each other in $z$. It follows that $X, Y$, and $Z$ are the midpoints of $U I_{a}, V I_{b}$ and $W I_{c}$, respectively, and that the triangles $U V W$, $X Y Z$ and $I_{a} I_{b} I_{c}$ are either translates of each other or homothetic with a common homothety centre.

Construct the points $T$ and $S$ such that the quadrilaterals $U V I W, X Y T Z$ and $I_{a} I_{b} S I_{c}$ are homothetic. Then $T$ is the midpoint of $I S$. Moreover, note that $\ell$ is the Simson line of the point $I$ with respect to the triangle $U V W$, hence $I$ belongs to the circumcircle of the triangle $U V W$, therefore $T$ belongs to $\Omega$.

Consider now the homothety or translation $h_{1}$ that maps $X Y Z T$ to $I_{a} I_{b} I_{c} S$ and the homothety $h_{2}$ with centre $I$ and factor $\frac{1}{2}$. Furthermore, let $h=h_{2} \circ h_{1}$. The transform $h$ can be a homothety or a translation, and

$$
h(T)=h_{2}\left(h_{1}(T)\right)=h_{2}(S)=T,
$$

hence $T$ is a fixed point of $h$. So, $h$ is a homothety with centre $T$. Note that $h_{2}$ maps the excentres $I_{a}, I_{b}, I_{c}$ to $X_{0}, Y_{0}, Z_{0}$ defined in the preamble. Thus the centre $T$ of the homothety taking $X Y Z$ to $X_{0} Y_{0} Z_{0}$ belongs to $\Omega$, and this completes the proof.

G6. A convex quadrilateral $A B C D$ satisfies $A B \cdot C D=B C \cdot D A$. A point $X$ is chosen inside the quadrilateral so that $\angle X A B=\angle X C D$ and $\angle X B C=\angle X D A$. Prove that $\angle A X B+$ $\angle C X D=180^{\circ}$.
(Poland)
Solution 1. Let $B^{\prime}$ be the reflection of $B$ in the internal angle bisector of $\angle A X C$, so that $\angle A X B^{\prime}=\angle C X B$ and $\angle C X B^{\prime}=\angle A X B$. If $X, D$, and $B^{\prime}$ are collinear, then we are done. Now assume the contrary.

On the ray $X B^{\prime}$ take a point $E$ such that $X E \cdot X B=X A \cdot X C$, so that $\triangle A X E \sim$ $\triangle B X C$ and $\triangle C X E \sim \triangle B X A$. We have $\angle X C E+\angle X C D=\angle X B A+\angle X A B<180^{\circ}$ and $\angle X A E+\angle X A D=\angle X D A+\angle X A D<180^{\circ}$, which proves that $X$ lies inside the angles $\angle E C D$ and $\angle E A D$ of the quadrilateral $E A D C$. Moreover, $X$ lies in the interior of exactly one of the two triangles $E A D, E C D$ (and in the exterior of the other).


The similarities mentioned above imply $X A \cdot B C=X B \cdot A E$ and $X B \cdot C E=X C \cdot A B$. Multiplying these equalities with the given equality $A B \cdot C D=B C \cdot D A$, we obtain $X A \cdot C D$. $C E=X C \cdot A D \cdot A E$, or, equivalently,

$$
\begin{equation*}
\frac{X A \cdot D E}{A D \cdot A E}=\frac{X C \cdot D E}{C D \cdot C E} . \tag{*}
\end{equation*}
$$

Lemma. Let $P Q R$ be a triangle, and let $X$ be a point in the interior of the angle $Q P R$ such that $\angle Q P X=\angle P R X$. Then $\frac{P X \cdot Q R}{P Q \cdot P R}<1$ if and only if $X$ lies in the interior of the triangle $P Q R$. Proof. The locus of points $X$ with $\angle Q P X=\angle P R X$ lying inside the angle $Q P R$ is an arc $\alpha$ of the circle $\gamma$ through $R$ tangent to $P Q$ at $P$. Let $\gamma$ intersect the line $Q R$ again at $Y$ (if $\gamma$ is tangent to $Q R$, then set $Y=R$ ). The similarity $\triangle Q P Y \sim \triangle Q R P$ yields $P Y=\frac{P Q \cdot P R}{Q R}$. Now it suffices to show that $P X<P Y$ if and only if $X$ lies in the interior of the triangle $P Q R$. Let $m$ be a line through $Y$ parallel to $P Q$. Notice that the points $Z$ of $\gamma$ satisfying $P Z<P Y$ are exactly those between the lines $m$ and $P Q$.

Case 1: $Y$ lies in the segment $Q R$ (see the left figure below).
In this case $Y$ splits $\alpha$ into two arcs $\overparen{P Y}$ and $\overparen{Y R}$. The arc $\overparen{P Y}$ lies inside the triangle $P Q R$, and $\widetilde{P Y}$ lies between $m$ and $P Q$, hence $P X<P Y$ for points $X \in \widehat{P Y}$. The other arc $\overline{Y R}$ lies outside triangle $P Q R$, and $\widehat{Y R}$ is on the opposite side of $m$ than $P$, hence $P X>P Y$ for $X \in \widehat{Y R}$.

Case 2: $Y$ lies on the ray $Q R$ beyond $R$ (see the right figure below).
In this case the whole arc $\alpha$ lies inside triangle $P Q R$, and between $m$ and $P Q$, thus $P X<$ $P Y$ for all $X \in \alpha$.


Applying the Lemma (to $\triangle E A D$ with the point $X$, and to $\triangle E C D$ with the point $X$ ), we obtain that exactly one of two expressions $\frac{X A \cdot D E}{A D \cdot A E}$ and $\frac{X C \cdot D E}{C D \cdot C E}$ is less than 1 , which contradicts (*).

Comment 1. One may show that $A B \cdot C D=X A \cdot X C+X B \cdot X D$. We know that $D, X, E$ are collinear and $\angle D C E=\angle C X D=180^{\circ}-\angle A X B$. Therefore,

$$
A B \cdot C D=X B \cdot \frac{\sin \angle A X B}{\sin \angle B A X} \cdot D E \cdot \frac{\sin \angle C E D}{\sin \angle D C E}=X B \cdot D E .
$$

Furthermore, $X B \cdot D E=X B \cdot(X D+X E)=X B \cdot X D+X B \cdot X E=X B \cdot X D+X A \cdot X C$.
Comment 2. For a convex quadrilateral $A B C D$ with $A B \cdot C D=B C \cdot D A$, it is known that $\angle D A C+\angle A B D+\angle B C A+\angle C D B=180^{\circ}$ (among other, it was used as a problem on the Regional round of All-Russian olympiad in 2012), but it seems that there is no essential connection between this fact and the original problem.

Solution 2. The solution consists of two parts. In Part 1 we show that it suffices to prove that

$$
\begin{equation*}
\frac{X B}{X D}=\frac{A B}{C D} \tag{1}
\end{equation*}
$$

and

$$
\begin{equation*}
\frac{X A}{X C}=\frac{D A}{B C} \tag{2}
\end{equation*}
$$

In Part 2 we establish these equalities.
Part 1. Using the sine law and applying (1) we obtain

$$
\frac{\sin \angle A X B}{\sin \angle X A B}=\frac{A B}{X B}=\frac{C D}{X D}=\frac{\sin \angle C X D}{\sin \angle X C D}
$$

so $\sin \angle A X B=\sin \angle C X D$ by the problem conditions. Similarly, (2) yields $\sin \angle D X A=$ $\sin \angle B X C$. If at least one of the pairs $(\angle A X B, \angle C X D)$ and $(\angle B X C, \angle D X A)$ consists of supplementary angles, then we are done. Otherwise, $\angle A X B=\angle C X D$ and $\angle D X A=\angle B X C$. In this case $X=A C \cap B D$, and the problem conditions yield that $A B C D$ is a parallelogram and hence a rhombus. In this last case the claim also holds.

Part 2. To prove the desired equality (1), invert $A B C D$ at centre $X$ with unit radius; the images of points are denoted by primes.

We have

$$
\angle A^{\prime} B^{\prime} C^{\prime}=\angle X B^{\prime} A^{\prime}+\angle X B^{\prime} C^{\prime}=\angle X A B+\angle X C B=\angle X C D+\angle X C B=\angle B C D .
$$

Similarly, the corresponding angles of quadrilaterals $A B C D$ and $D^{\prime} A^{\prime} B^{\prime} C^{\prime}$ are equal.
Moreover, we have

$$
A^{\prime} B^{\prime} \cdot C^{\prime} D^{\prime}=\frac{A B}{X A \cdot X B} \cdot \frac{C D}{X C \cdot X D}=\frac{B C}{X B \cdot X C} \cdot \frac{D A}{X D \cdot D A}=B^{\prime} C^{\prime} \cdot D^{\prime} A^{\prime}
$$



Now we need the following Lemma.
Lemma. Assume that the corresponding angles of convex quadrilaterals $X Y Z T$ and $X^{\prime} Y^{\prime} Z^{\prime} T^{\prime}$ are equal, and that $X Y \cdot Z T=Y Z \cdot T X$ and $X^{\prime} Y^{\prime} \cdot Z^{\prime} T^{\prime}=Y^{\prime} Z^{\prime} \cdot T^{\prime} X^{\prime}$. Then the two quadrilaterals are similar.
Proof. Take the quadrilateral $X Y Z_{1} T_{1}$ similar to $X^{\prime} Y^{\prime} Z^{\prime} T^{\prime}$ and sharing the side $X Y$ with $X Y Z T$, such that $Z_{1}$ and $T_{1}$ lie on the rays $Y Z$ and $X T$, respectively, and $Z_{1} T_{1} \| Z T$. We need to prove that $Z_{1}=Z$ and $T_{1}=T$. Assume the contrary. Without loss of generality, $T X>X T_{1}$. Let segments $X Z$ and $Z_{1} T_{1}$ intersect at $U$. We have

$$
\frac{T_{1} X}{T_{1} Z_{1}}<\frac{T_{1} X}{T_{1} U}=\frac{T X}{Z T}=\frac{X Y}{Y Z}<\frac{X Y}{Y Z_{1}},
$$

thus $T_{1} X \cdot Y Z_{1}<T_{1} Z_{1} \cdot X Y$. A contradiction.


It follows from the Lemma that the quadrilaterals $A B C D$ and $D^{\prime} A^{\prime} B^{\prime} C^{\prime}$ are similar, hence

$$
\frac{B C}{A B}=\frac{A^{\prime} B^{\prime}}{D^{\prime} A^{\prime}}=\frac{A B}{X A \cdot X B} \cdot \frac{X D \cdot X A}{D A}=\frac{A B}{A D} \cdot \frac{X D}{X B},
$$

and therefore

$$
\frac{X B}{X D}=\frac{A B^{2}}{B C \cdot A D}=\frac{A B^{2}}{A B \cdot C D}=\frac{A B}{C D} .
$$

We obtain (1), as desired; (2) is proved similarly.

Comment. Part 1 is an easy one, while part 2 seems to be crucial. On the other hand, after the proof of the similarity $D^{\prime} A^{\prime} B^{\prime} C^{\prime} \sim A B C D$ one may finish the solution in different ways, e.g., as follows. The similarity taking $D^{\prime} A^{\prime} B^{\prime} C^{\prime}$ to $A B C D$ maps $X$ to the point $X^{\prime}$ isogonally conjugate of $X$ with respect to $A B C D$ (i.e. to the point $X^{\prime}$ inside $A B C D$ such that $\angle B A X=\angle D A X^{\prime}$, $\left.\angle C B X=\angle A B X^{\prime}, \angle D C X=\angle B C X^{\prime}, \angle A D X=\angle C D X^{\prime}\right)$. It is known that the required equality $\angle A X B+\angle C X D=180^{\circ}$ is one of known conditions on a point $X$ inside $A B C D$ equivalent to the existence of its isogonal conjugate.

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G7.
Let $O$ be the circumcentre, and $\Omega$ be the circumcircle of an acute-angled triangle $A B C$.
Let $P$ be an arbitrary point on $\Omega$, distinct from $A, B, C$, and their antipodes in $\Omega$. Denote the circumcentres of the triangles $A O P, B O P$, and $C O P$ by $O_{A}, O_{B}$, and $O_{C}$, respectively. The lines $\ell_{A}, \ell_{B}$, and $\ell_{C}$ perpendicular to $B C, C A$, and $A B$ pass through $O_{A}, O_{B}$, and $O_{C}$, respectively. Prove that the circumcircle of the triangle formed by $\ell_{A}, \ell_{B}$, and $\ell_{C}$ is tangent to the line $O P$.
(Russia)
Solution. As usual, we denote the directed angle between the lines $a$ and $b$ by $\Varangle(a, b)$. We frequently use the fact that $a_{1} \perp a_{2}$ and $b_{1} \perp b_{2}$ yield $\Varangle\left(a_{1}, b_{1}\right)=\Varangle\left(a_{2}, b_{2}\right)$.

Let the lines $\ell_{B}$ and $\ell_{C}$ meet at $L_{A}$; define the points $L_{B}$ and $L_{C}$ similarly. Note that the sidelines of the triangle $L_{A} L_{B} L_{C}$ are perpendicular to the corresponding sidelines of $A B C$. Points $O_{A}, O_{B}, O_{C}$ are located on the corresponding sidelines of $L_{A} L_{B} L_{C}$; moreover, $O_{A}, O_{B}$, $O_{C}$ all lie on the perpendicular bisector of $O P$.


Claim 1. The points $L_{B}, P, O_{A}$, and $O_{C}$ are concyclic.
Proof. Since $O$ is symmetric to $P$ in $O_{A} O_{C}$, we have

$$
\Varangle\left(O_{A} P, O_{C} P\right)=\Varangle\left(O_{C} O, O_{A} O\right)=\Varangle(C P, A P)=\Varangle(C B, A B)=\Varangle\left(O_{A} L_{B}, O_{C} L_{B}\right) .
$$

Denote the circle through $L_{B}, P, O_{A}$, and $O_{C}$ by $\omega_{B}$. Define the circles $\omega_{A}$ and $\omega_{C}$ similarly. Claim 2. The circumcircle of the triangle $L_{A} L_{B} L_{C}$ passes through $P$.
Proof. From cyclic quadruples of points in the circles $\omega_{B}$ and $\omega_{C}$, we have

$$
\begin{aligned}
\Varangle\left(L_{C} L_{A}, L_{C} P\right) & =\Varangle\left(L_{C} O_{B}, L_{C} P\right)=\Varangle\left(O_{A} O_{B}, O_{A} P\right) \\
& =\Varangle\left(O_{A} O_{C}, O_{A} P\right)=\Varangle\left(L_{B} O_{C}, L_{B} P\right)=\Varangle\left(L_{B} L_{A}, L_{B} P\right) .
\end{aligned}
$$

Claim 3. The points $P, L_{C}$, and $C$ are collinear.
Proof. We have $\Varangle\left(P L_{C}, L_{C} L_{A}\right)=\Varangle\left(P L_{C}, L_{C} O_{B}\right)=\Varangle\left(P O_{A}, O_{A} O_{B}\right)$. Further, since $O_{A}$ is the centre of the circle $A O P, \Varangle\left(P O_{A}, O_{A} O_{B}\right)=\Varangle(P A, A O)$. As $O$ is the circumcentre of the triangle $P C A, \Varangle(P A, A O)=\pi / 2-\Varangle(C A, C P)=\Varangle\left(C P, L_{C} L_{A}\right)$. We obtain $\Varangle\left(P L_{C}, L_{C} L_{A}\right)=$ $\Varangle\left(C P, L_{C} L_{A}\right)$, which shows that $P \in C L_{C}$.

Similarly, the points $P, L_{A}, A$ are collinear, and the points $P, L_{B}, B$ are also collinear. Finally, the computation above also shows that

$$
\Varangle\left(O P, P L_{A}\right)=\Varangle(P A, A O)=\Varangle\left(P L_{C}, L_{C} L_{A}\right),
$$

which means that $O P$ is tangent to the circle $P L_{A} L_{B} L_{C}$.

Comment 1. The proof of Claim 2 may be replaced by the following remark: since $P$ belongs to the circles $\omega_{A}$ and $\omega_{C}, P$ is the Miquel point of the four lines $\ell_{A}, \ell_{B}, \ell_{C}$, and $O_{A} O_{B} O_{C}$.

Comment 2. Claims 2 and 3 can be proved in several different ways and, in particular, in the reverse order.

Claim 3 implies that the triangles $A B C$ and $L_{A} L_{B} L_{C}$ are perspective with perspector $P$. Claim 2 can be derived from this observation using spiral similarity. Consider the centre $Q$ of the spiral similarity that maps $A B C$ to $L_{A} L_{B} L_{C}$. From known spiral similarity properties, the points $L_{A}, L_{B}, P, Q$ are concyclic, and so are $L_{A}, L_{C}, P, Q$.

Comment 3. The final conclusion can also be proved it terms of spiral similarity: the spiral similarity with centre $Q$ located on the circle $A B C$ maps the circle $A B C$ to the circle $P L_{A} L_{B} L_{C}$. Thus these circles are orthogonal.

Comment 4. Notice that the homothety with centre $O$ and ratio 2 takes $O_{A}$ to $A^{\prime}$ that is the common point of tangents to $\Omega$ at $A$ and $P$. Similarly, let this homothety take $O_{B}$ to $B^{\prime}$ and $O_{C}$ to $C^{\prime}$. Let the tangents to $\Omega$ at $B$ and $C$ meet at $A^{\prime \prime}$, and define the points $B^{\prime \prime}$ and $C^{\prime \prime}$ similarly. Now, replacing labels $O$ with $I, \Omega$ with $\omega$, and swapping labels $A \leftrightarrow A^{\prime \prime}, B \leftrightarrow B^{\prime \prime}, C \leftrightarrow C^{\prime \prime}$ we obtain the following

Reformulation. Let $\omega$ be the incircle, and let $I$ be the incentre of a triangle $A B C$. Let $P$ be a point of $\omega$ (other than the points of contact of $\omega$ with the sides of $A B C$ ). The tangent to $\omega$ at $P$ meets the lines $A B, B C$, and $C A$ at $A^{\prime}, B^{\prime}$, and $C^{\prime}$, respectively. Line $\ell_{A}$ parallel to the internal angle bisector of $\angle B A C$ passes through $A^{\prime}$; define lines $\ell_{B}$ and $\ell_{C}$ similarly. Prove that the line $I P$ is tangent to the circumcircle of the triangle formed by $\ell_{A}, \ell_{B}$, and $\ell_{C}$.

Though this formulation is equivalent to the original one, it seems more challenging, since the point of contact is now "hidden".

## Number Theory

N1. Determine all pairs $(n, k)$ of distinct positive integers such that there exists a positive integer $s$ for which the numbers of divisors of $s n$ and of $s k$ are equal.
(Ukraine)
Answer: All pairs $(n, k)$ such that $n \nmid k$ and $k \nmid n$.
Solution. As usual, the number of divisors of a positive integer $n$ is denoted by $d(n)$. If $n=\prod_{i} p_{i}^{\alpha_{i}}$ is the prime factorisation of $n$, then $d(n)=\prod_{i}\left(\alpha_{i}+1\right)$.

We start by showing that one cannot find any suitable number $s$ if $k \mid n$ or $n \mid k$ (and $k \neq n$ ). Suppose that $n \mid k$, and choose any positive integer $s$. Then the set of divisors of $s n$ is a proper subset of that of $s k$, hence $d(s n)<d(s k)$. Therefore, the pair $(n, k)$ does not satisfy the problem requirements. The case $k \mid n$ is similar.

Now assume that $n \nmid k$ and $k \nmid n$. Let $p_{1}, \ldots, p_{t}$ be all primes dividing $n k$, and consider the prime factorisations

$$
n=\prod_{i=1}^{t} p_{i}^{\alpha_{i}} \quad \text { and } \quad k=\prod_{i=1}^{t} p_{i}^{\beta_{i}} .
$$

It is reasonable to search for the number $s$ having the form

$$
s=\prod_{i=1}^{t} p_{i}^{\gamma_{i}}
$$

The (nonnegative integer) exponents $\gamma_{i}$ should be chosen so as to satisfy

$$
\begin{equation*}
\frac{d(s n)}{d(s k)}=\prod_{i=1}^{t} \frac{\alpha_{i}+\gamma_{i}+1}{\beta_{i}+\gamma_{i}+1}=1 . \tag{1}
\end{equation*}
$$

First of all, if $\alpha_{i}=\beta_{i}$ for some $i$, then, regardless of the value of $\gamma_{i}$, the corresponding factor in (1) equals 1 and does not affect the product. So we may assume that there is no such index $i$. For the other factors in (1), the following lemma is useful.
Lemma. Let $\alpha>\beta$ be nonnegative integers. Then, for every integer $M \geqslant \beta+1$, there exists a nonnegative integer $\gamma$ such that

$$
\frac{\alpha+\gamma+1}{\beta+\gamma+1}=1+\frac{1}{M}=\frac{M+1}{M} .
$$

Proof.

$$
\frac{\alpha+\gamma+1}{\beta+\gamma+1}=1+\frac{1}{M} \Longleftrightarrow \frac{\alpha-\beta}{\beta+\gamma+1}=\frac{1}{M} \Longleftrightarrow \gamma=M(\alpha-\beta)-(\beta+1) \geqslant 0
$$

Now we can finish the solution. Without loss of generality, there exists an index $u$ such that $\alpha_{i}>\beta_{i}$ for $i=1,2, \ldots, u$, and $\alpha_{i}<\beta_{i}$ for $i=u+1, \ldots, t$. The conditions $n \nmid k$ and $k \nmid n$ mean that $1 \leqslant u \leqslant t-1$.

Choose an integer $X$ greater than all the $\alpha_{i}$ and $\beta_{i}$. By the lemma, we can define the numbers $\gamma_{i}$ so as to satisfy

$$
\begin{array}{ll}
\frac{\alpha_{i}+\gamma_{i}+1}{\beta_{i}+\gamma_{i}+1}=\frac{u X+i}{u X+i-1} & \text { for } i=1,2, \ldots, u, \text { and } \\
\frac{\beta_{u+i}+\gamma_{u+i}+1}{\alpha_{u+i}+\gamma_{u+i}+1}=\frac{(t-u) X+i}{(t-u) X+i-1} & \text { for } i=1,2, \ldots, t-u
\end{array}
$$

Then we will have

$$
\frac{d(s n)}{d(s k)}=\prod_{i=1}^{u} \frac{u X+i}{u X+i-1} \cdot \prod_{i=1}^{t-u} \frac{(t-u) X+i-1}{(t-u) X+i}=\frac{u(X+1)}{u X} \cdot \frac{(t-u) X}{(t-u)(X+1)}=1
$$

as required.
Comment. The lemma can be used in various ways, in order to provide a suitable value of $s$. In particular, one may apply induction on the number $t$ of prime factors, using identities like

$$
\frac{n}{n-1}=\frac{n^{2}}{n^{2}-1} \cdot \frac{n+1}{n} .
$$

N2. Let $n>1$ be a positive integer. Each cell of an $n \times n$ table contains an integer. Suppose that the following conditions are satisfied:
(i) Each number in the table is congruent to 1 modulo $n$;
(ii) The sum of numbers in any row, as well as the sum of numbers in any column, is congruent to $n$ modulo $n^{2}$.

Let $R_{i}$ be the product of the numbers in the $i^{\text {th }}$ row, and $C_{j}$ be the product of the numbers in the $j^{\text {th }}$ column. Prove that the sums $R_{1}+\cdots+R_{n}$ and $C_{1}+\cdots+C_{n}$ are congruent modulo $n^{4}$.
(Indonesia)
Solution 1. Let $A_{i, j}$ be the entry in the $i^{\text {th }}$ row and the $j^{\text {th }}$ column; let $P$ be the product of all $n^{2}$ entries. For convenience, denote $a_{i, j}=A_{i, j}-1$ and $r_{i}=R_{i}-1$. We show that

$$
\begin{equation*}
\sum_{i=1}^{n} R_{i} \equiv(n-1)+P \quad\left(\bmod n^{4}\right) \tag{1}
\end{equation*}
$$

Due to symmetry of the problem conditions, the sum of all the $C_{j}$ is also congruent to $(n-1)+P$ modulo $n^{4}$, whence the conclusion.

By condition $(i)$, the number $n$ divides $a_{i, j}$ for all $i$ and $j$. So, every product of at least two of the $a_{i, j}$ is divisible by $n^{2}$, hence
$R_{i}=\prod_{j=1}^{n}\left(1+a_{i, j}\right)=1+\sum_{j=1}^{n} a_{i, j}+\sum_{1 \leqslant j_{1}<j_{2} \leqslant n} a_{i, j_{1}} a_{i, j_{2}}+\cdots \equiv 1+\sum_{j=1}^{n} a_{i, j} \equiv 1-n+\sum_{j=1}^{n} A_{i, j} \quad\left(\bmod n^{2}\right)$
for every index $i$. Using condition (ii), we obtain $R_{i} \equiv 1\left(\bmod n^{2}\right)$, and so $n^{2} \mid r_{i}$.
Therefore, every product of at least two of the $r_{i}$ is divisible by $n^{4}$. Repeating the same argument, we obtain

$$
P=\prod_{i=1}^{n} R_{i}=\prod_{i=1}^{n}\left(1+r_{i}\right) \equiv 1+\sum_{i=1}^{n} r_{i} \quad\left(\bmod n^{4}\right)
$$

whence

$$
\sum_{i=1}^{n} R_{i}=n+\sum_{i=1}^{n} r_{i} \equiv n+(P-1) \quad\left(\bmod n^{4}\right)
$$

as desired.

Comment. The original version of the problem statement contained also the condition
(iii) The product of all the numbers in the table is congruent to 1 modulo $n^{4}$.

This condition appears to be superfluous, so it was omitted.
Solution 2. We present a more straightforward (though lengthier) way to establish (1). We also use the notation of $a_{i, j}$.

By condition ( $i$ ), all the $a_{i, j}$ are divisible by $n$. Therefore, we have

$$
\begin{aligned}
P=\prod_{i=1}^{n} \prod_{j=1}^{n}\left(1+a_{i, j}\right) \equiv 1+\sum_{(i, j)} a_{i, j} & +\sum_{\left(i_{1}, j_{1}\right),\left(i_{2}, j_{2}\right)} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}} \\
& +\sum_{\left(i_{1}, j_{1}\right),\left(i_{2}, j_{2}\right),\left(i_{3}, j_{3}\right)} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}} a_{i_{3}, j_{3}}\left(\bmod n^{4}\right),
\end{aligned}
$$

where the last two sums are taken over all unordered pairs/triples of pairwise different pairs $(i, j)$; such conventions are applied throughout the solution.

Similarly,

$$
\sum_{i=1}^{n} R_{i}=\sum_{i=1}^{n} \prod_{j=1}^{n}\left(1+a_{i, j}\right) \equiv n+\sum_{i} \sum_{j} a_{i, j}+\sum_{i} \sum_{j_{1}, j_{2}} a_{i, j_{1}} a_{i, j_{2}}+\sum_{i} \sum_{j_{1}, j_{2}, j_{3}} a_{i, j_{1}} a_{i, j_{2}} a_{i, j_{3}} \quad\left(\bmod n^{4}\right)
$$

Therefore,

$$
\begin{aligned}
P+(n-1)-\sum_{i} R_{i} \equiv \sum_{\substack{\left(i_{1}, j_{1}\right),\left(i_{2}, j_{2}\right) \\
i_{1} \neq i_{2}}} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}} & +\sum_{\substack{\left(i_{1}, j_{1}\right),\left(i_{2}, j_{2}\right),\left(i_{3}, j_{3}\right) \\
i_{1} \neq i_{2} \neq \neq i_{3} \neq i_{1}}} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}} a_{i_{3}, j_{3}} \\
& +\sum_{\substack{\left(i_{1}, j_{1}\right),\left(i_{2}, j_{2}\right),\left(i_{3}, j_{3}\right) \\
i_{1} \neq i_{2}=i_{3}}} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}} a_{i_{3}, j_{3}}\left(\bmod n^{4}\right) .
\end{aligned}
$$

We show that in fact each of the three sums appearing in the right-hand part of this congruence is divisible by $n^{4}$; this yields (1). Denote those three sums by $\Sigma_{1}, \Sigma_{2}$, and $\Sigma_{3}$ in order of appearance. Recall that by condition (ii) we have

$$
\sum_{j} a_{i, j} \equiv 0 \quad\left(\bmod n^{2}\right) \quad \text { for all indices } i .
$$

For every two indices $i_{1}<i_{2}$ we have

$$
\sum_{j_{1}} \sum_{j_{2}} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}}=\left(\sum_{j_{1}} a_{i_{1}, j_{1}}\right) \cdot\left(\sum_{j_{2}} a_{i_{2}, j_{2}}\right) \equiv 0 \quad\left(\bmod n^{4}\right),
$$

since each of the two factors is divisible by $n^{2}$. Summing over all pairs $\left(i_{1}, i_{2}\right)$ we obtain $n^{4} \mid \Sigma_{1}$.
Similarly, for every three indices $i_{1}<i_{2}<i_{3}$ we have

$$
\sum_{j_{1}} \sum_{j_{2}} \sum_{j_{3}} a_{i_{1}, j_{1}} a_{i_{2}, j_{2}} a_{i_{3}, j_{3}}=\left(\sum_{j_{1}} a_{i_{1}, j_{1}}\right) \cdot\left(\sum_{j_{2}} a_{i_{2}, j_{2}}\right) \cdot\left(\sum_{j_{3}} a_{i_{3}, j_{3}}\right)
$$

which is divisible even by $n^{6}$. Hence $n^{4} \mid \Sigma_{2}$.
Finally, for every indices $i_{1} \neq i_{2}=i_{3}$ and $j_{2}<j_{3}$ we have

$$
a_{i_{2}, j_{2}} \cdot a_{i_{2}, j_{3}} \cdot \sum_{j_{1}} a_{i_{1}, j_{1}} \equiv 0 \quad\left(\bmod n^{4}\right),
$$

since the three factors are divisible by $n, n$, and $n^{2}$, respectively. Summing over all 4 -tuples of indices $\left(i_{1}, i_{2}, j_{2}, j_{3}\right)$ we get $n^{4} \mid \Sigma_{3}$.

N3. Define the sequence $a_{0}, a_{1}, a_{2}, \ldots$ by $a_{n}=2^{n}+2^{\lfloor n / 2\rfloor}$. Prove that there are infinitely many terms of the sequence which can be expressed as a sum of (two or more) distinct terms of the sequence, as well as infinitely many of those which cannot be expressed in such a way.

Solution 1. Call a nonnegative integer representable if it equals the sum of several (possibly 0 or 1) distinct terms of the sequence. We say that two nonnegative integers $b$ and $c$ are equivalent (written as $b \sim c$ ) if they are either both representable or both non-representable.

One can easily compute

$$
S_{n-1}:=a_{0}+\cdots+a_{n-1}=2^{n}+2^{[n / 2]}+2^{[n / 2]}-3 .
$$

Indeed, we have $S_{n}-S_{n-1}=2^{n}+2^{\lfloor n / 2\rfloor}=a_{n}$ so we can use the induction. In particular, $S_{2 k-1}=2^{2 k}+2^{k+1}-3$.

Note that, if $n \geqslant 3$, then $2^{[n / 2]} \geqslant 2^{2}>3$, so

$$
S_{n-1}=2^{n}+2^{[n / 2]}+2^{[n / 2]}-3>2^{n}+2^{[n / 2]}=a_{n} .
$$

Also notice that $S_{n-1}-a_{n}=2^{[n / 2]}-3<a_{n}$.
The main tool of the solution is the following claim.
Claim 1. Assume that $b$ is a positive integer such that $S_{n-1}-a_{n}<b<a_{n}$ for some $n \geqslant 3$. Then $b \sim S_{n-1}-b$.
Proof. As seen above, we have $S_{n-1}>a_{n}$. Denote $c=S_{n-1}-b$; then $S_{n-1}-a_{n}<c<a_{n}$, so the roles of $b$ and $c$ are symmetrical.

Assume that $b$ is representable. The representation cannot contain $a_{i}$ with $i \geqslant n$, since $b<a_{n}$. So $b$ is the sum of some subset of $\left\{a_{0}, a_{1}, \ldots, a_{n-1}\right\}$; then $c$ is the sum of the complement. The converse is obtained by swapping $b$ and $c$.

We also need the following version of this claim.
Claim 2. For any $n \geqslant 3$, the number $a_{n}$ can be represented as a sum of two or more distinct terms of the sequence if and only if $S_{n-1}-a_{n}=2^{[n / 2]}-3$ is representable.
Proof. Denote $c=S_{n-1}-a_{n}<a_{n}$. If $a_{n}$ satisfies the required condition, then it is the sum of some subset of $\left\{a_{0}, a_{1}, \ldots, a_{n-1}\right\}$; then $c$ is the sum of the complement. Conversely, if $c$ is representable, then its representation consists only of the numbers from $\left\{a_{0}, \ldots, a_{n-1}\right\}$, so $a_{n}$ is the sum of the complement.

By Claim 2, in order to prove the problem statement, it suffices to find infinitely many representable numbers of the form $2^{t}-3$, as well as infinitely many non-representable ones.
Claim 3. For every $t \geqslant 3$, we have $2^{t}-3 \sim 2^{4 t-6}-3$, and $2^{4 t-6}-3>2^{t}-3$.
Proof. The inequality follows from $t \geqslant 3$. In order to prove the equivalence, we apply Claim 1 twice in the following manner.

First, since $S_{2 t-3}-a_{2 t-2}=2^{t-1}-3<2^{t}-3<2^{2 t-2}+2^{t-1}=a_{2 t-2}$, by Claim 1 we have $2^{t}-3 \sim S_{2 t-3}-\left(2^{t}-3\right)=2^{2 t-2}$.

Second, since $S_{4 t-7}-a_{4 t-6}=2^{2 t-3}-3<2^{2 t-2}<2^{4 t-6}+2^{2 t-3}=a_{4 t-6}$, by Claim 1 we have $2^{2 t-2} \sim S_{4 t-7}-2^{2 t-2}=2^{4 t-6}-3$.

Therefore, $2^{t}-3 \sim 2^{2 t-2} \sim 2^{4 t-6}-3$, as required.
Now it is easy to find the required numbers. Indeed, the number $2^{3}-3=5=a_{0}+a_{1}$ is representable, so Claim 3 provides an infinite sequence of representable numbers

$$
2^{3}-3 \sim 2^{6}-3 \sim 2^{18}-3 \sim \cdots \sim 2^{t}-3 \sim 2^{4 t-6}-3 \sim \cdots .
$$

On the other hand, the number $2^{7}-3=125$ is non-representable (since by Claim 1 we have $125 \sim S_{6}-125=24 \sim S_{4}-24=17 \sim S_{3}-17=4$ which is clearly non-representable). So Claim 3 provides an infinite sequence of non-representable numbers

$$
2^{7}-3 \sim 2^{22}-3 \sim 2^{82}-3 \sim \cdots \sim 2^{t}-3 \sim 2^{4 t-6}-3 \sim \cdots
$$

Solution 2. We keep the notion of representability and the notation $S_{n}$ from the previous solution. We say that an index $n$ is good if $a_{n}$ writes as a sum of smaller terms from the sequence $a_{0}, a_{1}, \ldots$. Otherwise we say it is bad. We must prove that there are infinitely many good indices, as well as infinitely many bad ones.
Lemma 1. If $m \geqslant 0$ is an integer, then $4^{m}$ is representable if and only if either of $2 m+1$ and $2 m+2$ is good.
Proof. The case $m=0$ is obvious, so we may assume that $m \geqslant 1$. Let $n=2 m+1$ or $2 m+2$. Then $n \geqslant 3$. We notice that

$$
S_{n-1}<a_{n-2}+a_{n} .
$$

The inequality writes as $2^{n}+2^{[n / 2]}+2^{\lfloor n / 2\rfloor}-3<2^{n}+2^{\lfloor n / 2\rfloor}+2^{n-2}+2^{\lfloor n / 2\rfloor-1}$, i.e. as $2^{[n / 2\rceil}<$ $2^{n-2}+2^{\lfloor n / 2\rfloor-1}+3$. If $n \geqslant 4$, then $n / 2 \leqslant n-2$, so $\lceil n / 2\rceil \leqslant n-2$ and $2^{[n / 2\rceil} \leqslant 2^{n-2}$. For $n=3$ the inequality verifies separately.

If $n$ is good, then $a_{n}$ writes as $a_{n}=a_{i_{1}}+\cdots+a_{i_{r}}$, where $r \geqslant 2$ and $i_{1}<\cdots<i_{r}<n$. Then $i_{r}=n-1$ and $i_{r-1}=n-2$, for if $n-1$ or $n-2$ is missing from the sequence $i_{1}, \ldots, i_{r}$, then $a_{i_{1}}+\cdots+a_{i_{r}} \leqslant a_{0}+\cdots+a_{n-3}+a_{n-1}=S_{n-1}-a_{n-2}<a_{n}$. Thus, if $n$ is good, then both $a_{n}-a_{n-1}$ and $a_{n}-a_{n-1}-a_{n-2}$ are representable.

We now consider the cases $n=2 m+1$ and $n=2 m+2$ separately.
If $n=2 m+1$, then $a_{n}-a_{n-1}=a_{2 m+1}-a_{2 m}=\left(2^{2 m+1}+2^{m}\right)-\left(2^{2 m}+2^{m}\right)=2^{2 m}$. So we proved that, if $2 m+1$ is good, then $2^{2 m}$ is representable. Conversely, if $2^{2 m}$ is representable, then $2^{2 m}<a_{2 m}$, so $2^{2 m}$ is a sum of some distinct terms $a_{i}$ with $i<2 m$. It follows that $a_{2 m+1}=a_{2 m}+2^{2 m}$ writes as $a_{2 m}$ plus a sum of some distinct terms $a_{i}$ with $i<2 m$. Hence $2 m+1$ is good.

If $n=2 m+2$, then $a_{n}-a_{n-1}-a_{n-2}=a_{2 m+2}-a_{2 m+1}-a_{2 m}=\left(2^{2 m+2}+2^{m+1}\right)-\left(2^{2 m+1}+\right.$ $\left.2^{m}\right)-\left(2^{2 m}+2^{m}\right)=2^{2 m}$. So we proved that, if $2 m+2$ is good, then $2^{2 m}$ is representable. Conversely, if $2^{2 m}$ is representable, then, as seen in the previous case, it writes as a sum of some distinct terms $a_{i}$ with $i<2 m$. Hence $a_{2 m+2}=a_{2 m+1}+a_{2 m}+2^{2 m}$ writes as $a_{2 m+1}+a_{2 m}$ plus a sum of some distinct terms $a_{i}$ with $i<2 m$. Thus $2 m+2$ is good.

Lemma 2. If $k \geqslant 2$, then $2^{4 k-2}$ is representable if and only if $2^{k+1}$ is representable.
In particular, if $s \geqslant 2$, then $4^{s}$ is representable if and only if $4^{4 s-3}$ is representable. Also, $4^{4 s-3}>4^{s}$.
Proof. We have $2^{4 k-2}<a_{4 k-2}$, so in a representation of $2^{4 k-2}$ we can have only terms $a_{i}$ with $i \leqslant 4 k-3$. Notice that

$$
a_{0}+\cdots+a_{4 k-3}=2^{4 k-2}+2^{2 k}-3<2^{4 k-2}+2^{2 k}+2^{k}=2^{4 k-2}+a_{2 k}
$$

Hence, any representation of $2^{4 k-2}$ must contain all terms from $a_{2 k}$ to $a_{4 k-3}$. (If any of these terms is missing, then the sum of the remaining ones is $\leqslant\left(a_{0}+\cdots+a_{4 k-3}\right)-a_{2 k}<2^{4 k-2}$.) Hence, if $2^{4 k-2}$ is representable, then $2^{4 k-2}-\sum_{i=2 k}^{4 k-3} a_{i}$ is representable. But
$2^{4 k-2}-\sum_{i=2 k}^{4 k-3} a_{i}=2^{4 k-2}-\left(S_{4 k-3}-S_{2 k-1}\right)=2^{4 k-2}-\left(2^{4 k-2}+2^{2 k}-3\right)+\left(2^{2 k}+2^{k+1}-3\right)=2^{k+1}$.
So, if $2^{4 k-2}$ is representable, then $2^{k+1}$ is representable. Conversely, if $2^{k+1}$ is representable, then $2^{k+1}<2^{2 k}+2^{k}=a_{2 k}$, so $2^{k+1}$ writes as a sum of some distinct terms $a_{i}$ with $i<2 k$. It follows that $2^{4 k-2}=\sum_{i=2 k}^{4 k-3} a_{i}+2^{k+1}$ writes as $a_{4 k-3}+a_{4 k-4}+\cdots+a_{2 k}$ plus the sum of some distinct terms $a_{i}$ with $i<2 k$. Hence $2^{4 k-2}$ is representable.

For the second statement, if $s \geqslant 2$, then we just take $k=2 s-1$ and we notice that $2^{k+1}=4^{s}$ and $2^{4 k-2}=4^{4 s-3}$. Also, $s \geqslant 2$ implies that $4 s-3>s$.

Now $4^{2}=a_{2}+a_{3}$ is representable, whereas $4^{6}=4096$ is not. Indeed, note that $4^{6}=2^{12}<a_{12}$, so the only available terms for a representation are $a_{0}, \ldots, a_{11}$, i.e., $2,3,6,10,20,36,72$, $136,272,528,1056,2080$. Their sum is $S_{11}=4221$, which exceeds 4096 by 125. Then any representation of 4096 must contain all the terms from $a_{0}, \ldots, a_{11}$ that are greater that 125 , i.e., $136,272,528,1056,2080$. Their sum is 4072 . Since $4096-4072=24$ and 24 is clearly not representable, 4096 is non-representable as well.

Starting with these values of $m$, by using Lemma 2, we can obtain infinitely many representable powers of 4 , as well as infinitely many non-representable ones. By Lemma 1 , this solves our problem.

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N4. Let $a_{1}, a_{2}, \ldots, a_{n}, \ldots$ be a sequence of positive integers such that

$$
\frac{a_{1}}{a_{2}}+\frac{a_{2}}{a_{3}}+\cdots+\frac{a_{n-1}}{a_{n}}+\frac{a_{n}}{a_{1}}
$$

is an integer for all $n \geqslant k$, where $k$ is some positive integer. Prove that there exists a positive integer $m$ such that $a_{n}=a_{n+1}$ for all $n \geqslant m$.
(Mongolia)
Solution 1. The argument hinges on the following two facts: Let $a, b, c$ be positive integers such that $N=b / c+(c-b) / a$ is an integer.
(1) If $\operatorname{gcd}(a, c)=1$, then $c$ divides $b$; and
(2) If $\operatorname{gcd}(a, b, c)=1$, then $\operatorname{gcd}(a, b)=1$.

To prove (1), write $a b=c(a N+b-c)$. Since $\operatorname{gcd}(a, c)=1$, it follows that $c$ divides $b$. To prove (2), write $c^{2}-b c=a(c N-b)$ to infer that $a$ divides $c^{2}-b c$. Letting $d=\operatorname{gcd}(a, b)$, it follows that $d$ divides $c^{2}$, and since the two are relatively prime by hypothesis, $d=1$.

Now, let $s_{n}=a_{1} / a_{2}+a_{2} / a_{3}+\cdots+a_{n-1} / a_{n}+a_{n} / a_{1}$, let $\delta_{n}=\operatorname{gcd}\left(a_{1}, a_{n}, a_{n+1}\right)$ and write

$$
s_{n+1}-s_{n}=\frac{a_{n}}{a_{n+1}}+\frac{a_{n+1}-a_{n}}{a_{1}}=\frac{a_{n} / \delta_{n}}{a_{n+1} / \delta_{n}}+\frac{a_{n+1} / \delta_{n}-a_{n} / \delta_{n}}{a_{1} / \delta_{n}} .
$$

Let $n \geqslant k$. Since $\operatorname{gcd}\left(a_{1} / \delta_{n}, a_{n} / \delta_{n}, a_{n+1} / \delta_{n}\right)=1$, it follows by (2) that $\operatorname{gcd}\left(a_{1} / \delta_{n}, a_{n} / \delta_{n}\right)=1$. Let $d_{n}=\operatorname{gcd}\left(a_{1}, a_{n}\right)$. Then $d_{n}=\delta_{n} \cdot \operatorname{gcd}\left(a_{1} / \delta_{n}, a_{n} / \delta_{n}\right)=\delta_{n}$, so $d_{n}$ divides $a_{n+1}$, and therefore $d_{n}$ divides $d_{n+1}$.

Consequently, from some rank on, the $d_{n}$ form a nondecreasing sequence of integers not exceeding $a_{1}$, so $d_{n}=d$ for all $n \geqslant \ell$, where $\ell$ is some positive integer.

Finally, since $\operatorname{gcd}\left(a_{1} / d, a_{n+1} / d\right)=1$, it follows by (1) that $a_{n+1} / d$ divides $a_{n} / d$, so $a_{n} \geqslant a_{n+1}$ for all $n \geqslant \ell$. The conclusion follows.

Solution 2. We use the same notation $s_{n}$. This time, we explore the exponents of primes in the prime factorizations of the $a_{n}$ for $n \geqslant k$.

To start, for every $n \geqslant k$, we know that the number

$$
\begin{equation*}
s_{n+1}-s_{n}=\frac{a_{n}}{a_{n+1}}+\frac{a_{n+1}}{a_{1}}-\frac{a_{n}}{a_{1}} \tag{*}
\end{equation*}
$$

is integer. Multiplying it by $a_{1}$ we obtain that $a_{1} a_{n} / a_{n+1}$ is integer as well, so that $a_{n+1} \mid a_{1} a_{n}$. This means that $a_{n} \mid a_{1}^{n-k} a_{k}$, so all prime divisors of $a_{n}$ are among those of $a_{1} a_{k}$. There are finitely many such primes; therefore, it suffices to prove that the exponent of each of them in the prime factorization of $a_{n}$ is eventually constant.

Choose any prime $p \mid a_{1} a_{k}$. Recall that $v_{p}(q)$ is the standard notation for the exponent of $p$ in the prime factorization of a nonzero rational number $q$. Say that an index $n \geqslant k$ is large if $v_{p}\left(a_{n}\right) \geqslant v_{p}\left(a_{1}\right)$. We separate two cases.
Case 1: There exists a large index $n$.
If $v_{p}\left(a_{n+1}\right)<v_{p}\left(a_{1}\right)$, then $v_{p}\left(a_{n} / a_{n+1}\right)$ and $v_{p}\left(a_{n} / a_{1}\right)$ are nonnegative, while $v_{p}\left(a_{n+1} / a_{1}\right)<0$; hence (*) cannot be an integer. This contradiction shows that index $n+1$ is also large.

On the other hand, if $v_{p}\left(a_{n+1}\right)>v_{p}\left(a_{n}\right)$, then $v_{p}\left(a_{n} / a_{n+1}\right)<0$, while $v_{p}\left(\left(a_{n+1}-a_{n}\right) / a_{1}\right) \geqslant 0$, so (*) is not integer again. Thus, $v_{p}\left(a_{1}\right) \leqslant v_{p}\left(a_{n+1}\right) \leqslant v_{p}\left(a_{n}\right)$.

The above arguments can now be applied successively to indices $n+1, n+2, \ldots$, showing that all the indices greater than $n$ are large, and the sequence $v_{p}\left(a_{n}\right), v_{p}\left(a_{n+1}\right), v_{p}\left(a_{n+2}\right), \ldots$ is nonincreasing - hence eventually constant.

Case 2: There is no large index.
We have $v_{p}\left(a_{1}\right)>v_{p}\left(a_{n}\right)$ for all $n \geqslant k$. If we had $v_{p}\left(a_{n+1}\right)<v_{p}\left(a_{n}\right)$ for some $n \geqslant k$, then $v_{p}\left(a_{n+1} / a_{1}\right)<v_{p}\left(a_{n} / a_{1}\right)<0<v_{p}\left(a_{n} / a_{n+1}\right)$ which would also yield that $(*)$ is not integer. Therefore, in this case the sequence $v_{p}\left(a_{k}\right), v_{p}\left(a_{k+1}\right), v_{p}\left(a_{k+2}\right), \ldots$ is nondecreasing and bounded by $v_{p}\left(a_{1}\right)$ from above; hence it is also eventually constant.

Comment. Given any positive odd integer $m$, consider the $m$-tuple $\left(2,2^{2}, \ldots, 2^{m-1}, 2^{m}\right)$. Appending an infinite string of 1's to this $m$-tuple yields an eventually constant sequence of integers satisfying the condition in the statement, and shows that the rank from which the sequence stabilises may be arbitrarily large.

There are more sophisticated examples. The solution to part (b) of 10532, Amer. Math. Monthly, Vol. 105 No. 8 (Oct. 1998), 775-777 (available at https://www.jstor.org/stable/2589009), shows that, for every integer $m \geqslant 5$, there exists an $m$-tuple ( $a_{1}, a_{2}, \ldots, a_{m}$ ) of pairwise distinct positive integers such that $\operatorname{gcd}\left(a_{1}, a_{2}\right)=\operatorname{gcd}\left(a_{2}, a_{3}\right)=\cdots=\operatorname{gcd}\left(a_{m-1}, a_{m}\right)=\operatorname{gcd}\left(a_{m}, a_{1}\right)=1$, and the sum $a_{1} / a_{2}+a_{2} / a_{3}+\cdots+a_{m-1} / a_{m}+a_{m} / a_{1}$ is an integer. Letting $a_{m+k}=a_{1}, k=1,2, \ldots$, extends such an $m$-tuple to an eventually constant sequence of positive integers satisfying the condition in the statement of the problem at hand.

Here is the example given by the proposers of 10532. Let $b_{1}=2$, let $b_{k+1}=1+b_{1} \cdots b_{k}=$ $1+b_{k}\left(b_{k}-1\right), k \geqslant 1$, and set $B_{m}=b_{1} \cdots b_{m-4}=b_{m-3}-1$. The $m$-tuple $\left(a_{1}, a_{2}, \ldots, a_{m}\right)$ defined below satisfies the required conditions:

$$
\begin{aligned}
a_{1}=1, & a_{2}=\left(8 B_{m}+1\right) B_{m}+8, \quad a_{3}=8 B_{m}+1, \quad a_{k}=b_{m-k} \quad \text { for } 4 \leqslant k \leqslant m-1, \\
& a_{m}=\frac{a_{2}}{2} \cdot a_{3} \cdot \frac{B_{m}}{2}=\left(\frac{1}{2}\left(8 B_{m}+1\right) B_{m}+4\right) \cdot\left(8 B_{m}+1\right) \cdot \frac{B_{m}}{2} .
\end{aligned}
$$

It is readily checked that $a_{1}<a_{m-1}<a_{m-2}<\cdots<a_{3}<a_{2}<a_{m}$. For further details we refer to the solution mentioned above. Acquaintance with this example (or more elaborated examples derived from) offers no advantage in tackling the problem.

N5. Four positive integers $x, y, z$, and $t$ satisfy the relations

$$
\begin{equation*}
x y-z t=x+y=z+t . \tag{*}
\end{equation*}
$$

Is it possible that both $x y$ and $z t$ are perfect squares?
(Russia)
Answer: No.
Solution 1. Arguing indirectly, assume that $x y=a^{2}$ and $z t=c^{2}$ with $a, c>0$.
Suppose that the number $x+y=z+t$ is odd. Then $x$ and $y$ have opposite parity, as well as $z$ and $t$. This means that both $x y$ and $z t$ are even, as well as $x y-z t=x+y$; a contradiction. Thus, $x+y$ is even, so the number $s=\frac{x+y}{2}=\frac{z+t}{2}$ is a positive integer.

Next, we set $b=\frac{|x-y|}{2}, d=\frac{|z-t|}{2}$. Now the problem conditions yield

$$
\begin{equation*}
s^{2}=a^{2}+b^{2}=c^{2}+d^{2} \tag{1}
\end{equation*}
$$

and

$$
\begin{equation*}
2 s=a^{2}-c^{2}=d^{2}-b^{2} \tag{2}
\end{equation*}
$$

(the last equality in (2) follows from (1)). We readily get from (2) that $a, d>0$.
In the sequel we will use only the relations (1) and (2), along with the fact that $a, d, s$ are positive integers, while $b$ and $c$ are nonnegative integers, at most one of which may be zero. Since both relations are symmetric with respect to the simultaneous swappings $a \leftrightarrow d$ and $b \leftrightarrow c$, we assume, without loss of generality, that $b \geqslant c$ (and hence $b>0$ ). Therefore, $d^{2}=2 s+b^{2}>c^{2}$, whence

$$
\begin{equation*}
d^{2}>\frac{c^{2}+d^{2}}{2}=\frac{s^{2}}{2} . \tag{3}
\end{equation*}
$$

On the other hand, since $d^{2}-b^{2}$ is even by (2), the numbers $b$ and $d$ have the same parity, so $0<b \leqslant d-2$. Therefore,

$$
\begin{equation*}
2 s=d^{2}-b^{2} \geqslant d^{2}-(d-2)^{2}=4(d-1), \quad \text { i.e., } \quad d \leqslant \frac{s}{2}+1 . \tag{4}
\end{equation*}
$$

Combining (3) and (4) we obtain

$$
2 s^{2}<4 d^{2} \leqslant 4\left(\frac{s}{2}+1\right)^{2}, \quad \text { or } \quad(s-2)^{2}<8
$$

which yields $s \leqslant 4$.
Finally, an easy check shows that each number of the form $s^{2}$ with $1 \leqslant s \leqslant 4$ has a unique representation as a sum of two squares, namely $s^{2}=s^{2}+0^{2}$. Thus, (1) along with $a, d>0$ imply $b=c=0$, which is impossible.

Solution 2. We start with a complete description of all 4-tuples ( $x, y, z, t$ ) of positive integers satisfying (*). As in the solution above, we notice that the numbers

$$
s=\frac{x+y}{2}=\frac{z+t}{2}, \quad p=\frac{x-y}{2}, \quad \text { and } \quad q=\frac{z-t}{2}
$$

are integers (we may, and will, assume that $p, q \geqslant 0$ ). We have

$$
2 s=x y-z t=(s+p)(s-p)-(s+q)(s-q)=q^{2}-p^{2},
$$

so $p$ and $q$ have the same parity, and $q>p$.

Set now $k=\frac{q-p}{2}, \ell=\frac{q+p}{2}$. Then we have $s=\frac{q^{2}-p^{2}}{2}=2 k \ell$ and hence

$$
\begin{array}{ll}
x=s+p=2 k \ell-k+\ell, & y=s-p=2 k \ell+k-\ell \\
z=s+q=2 k \ell+k+\ell, & t=s-q=2 k \ell-k-\ell . \tag{5}
\end{array}
$$

Recall here that $\ell \geqslant k>0$ and, moreover, $(k, \ell) \neq(1,1)$, since otherwise $t=0$.
Assume now that both $x y$ and $z t$ are squares. Then $x y z t$ is also a square. On the other hand, we have

$$
\begin{align*}
x y z t=(2 k \ell-k+\ell) & (2 k \ell+k-\ell)(2 k \ell+k+\ell)(2 k \ell-k-\ell) \\
& =\left(4 k^{2} \ell^{2}-(k-\ell)^{2}\right)\left(4 k^{2} \ell^{2}-(k+\ell)^{2}\right)=\left(4 k^{2} \ell^{2}-k^{2}-\ell^{2}\right)^{2}-4 k^{2} \ell^{2} . \tag{6}
\end{align*}
$$

Denote $D=4 k^{2} \ell^{2}-k^{2}-\ell^{2}>0$. From (6) we get $D^{2}>x y z t$. On the other hand,

$$
\begin{aligned}
&(D-1)^{2}=D^{2}-2\left(4 k^{2} \ell^{2}-k^{2}-\ell^{2}\right)+1=\left(D^{2}-4 k^{2} \ell^{2}\right)-\left(2 k^{2}-1\right)\left(2 \ell^{2}-1\right)+2 \\
&=x y z t-\left(2 k^{2}-1\right)\left(2 \ell^{2}-1\right)+2<x y z t
\end{aligned}
$$

since $\ell \geqslant 2$ and $k \geqslant 1$. Thus $(D-1)^{2}<x y z t<D^{2}$, and xyzt cannot be a perfect square; a contradiction.

Comment. The first part of Solution 2 shows that all 4 -tuples of positive integers $x \geqslant y, z \geqslant t$ satisfying (*) have the form (5), where $\ell \geqslant k>0$ and $\ell \geqslant 2$. The converse is also true: every pair of positive integers $\ell \geqslant k>0$, except for the pair $k=\ell=1$, generates via (5) a 4 -tuple of positive integers satisfying (*).

N6. Let $f:\{1,2,3, \ldots\} \rightarrow\{2,3, \ldots\}$ be a function such that $f(m+n) \mid f(m)+f(n)$ for all pairs $m, n$ of positive integers. Prove that there exists a positive integer $c>1$ which divides all values of $f$.
(Mexico)
Solution 1. For every positive integer $m$, define $S_{m}=\{n: m \mid f(n)\}$.
Lemma. If the set $S_{m}$ is infinite, then $S_{m}=\{d, 2 d, 3 d, \ldots\}=d \cdot \mathbb{Z}_{>0}$ for some positive integer $d$. Proof. Let $d=\min S_{m}$; the definition of $S_{m}$ yields $m \mid f(d)$.

Whenever $n \in S_{m}$ and $n>d$, we have $m|f(n)| f(n-d)+f(d)$, so $m \mid f(n-d)$ and therefore $n-d \in S_{m}$. Let $r \leqslant d$ be the least positive integer with $n \equiv r(\bmod d)$; repeating the same step, we can see that $n-d, n-2 d, \ldots, r \in S_{m}$. By the minimality of $d$, this shows $r=d$ and therefore $d \mid n$.

Starting from an arbitrarily large element of $S_{m}$, the process above reaches all multiples of $d$; so they all are elements of $S_{m}$.

The solution for the problem will be split into two cases.

## Case 1: The function $f$ is bounded.

Call a prime $p$ frequent if the set $S_{p}$ is infinite, i.e., if $p$ divides $f(n)$ for infinitely many positive integers $n$; otherwise call $p$ sporadic. Since the function $f$ is bounded, there are only a finite number of primes that divide at least one $f(n)$; so altogether there are finitely many numbers $n$ such that $f(n)$ has a sporadic prime divisor. Let $N$ be a positive integer, greater than all those numbers $n$.

Let $p_{1}, \ldots, p_{k}$ be the frequent primes. By the lemma we have $S_{p_{i}}=d_{i} \cdot \mathbb{Z}_{>0}$ for some $d_{i}$. Consider the number

$$
n=N d_{1} d_{2} \cdots d_{k}+1
$$

Due to $n>N$, all prime divisors of $f(n)$ are frequent primes. Let $p_{i}$ be any frequent prime divisor of $f(n)$. Then $n \in S_{p_{i}}$, and therefore $d_{i} \mid n$. But $n \equiv 1\left(\bmod d_{i}\right)$, which means $d_{i}=1$. Hence $S_{p_{i}}=1 \cdot \mathbb{Z}_{>0}=\mathbb{Z}_{>0}$ and therefore $p_{i}$ is a common divisor of all values $f(n)$.

## Case 2: $f$ is unbounded.

We prove that $f(1)$ divides all $f(n)$.
Let $a=f(1)$. Since $1 \in S_{a}$, by the lemma it suffices to prove that $S_{a}$ is an infinite set.
Call a positive integer $p$ a peak if $f(p)>\max (f(1), \ldots, f(p-1))$. Since $f$ is not bounded, there are infinitely many peaks. Let $1=p_{1}<p_{2}<\ldots$ be the sequence of all peaks, and let $h_{k}=f\left(p_{k}\right)$. Notice that for any peak $p_{i}$ and for any $k<p_{i}$, we have $f\left(p_{i}\right) \mid f(k)+f\left(p_{i}-k\right)<$ $2 f\left(p_{i}\right)$, hence

$$
\begin{equation*}
f(k)+f\left(p_{i}-k\right)=f\left(p_{i}\right)=h_{i} . \tag{1}
\end{equation*}
$$

By the pigeonhole principle, among the numbers $h_{1}, h_{2}, \ldots$ there are infinitely many that are congruent modulo $a$. Let $k_{0}<k_{1}<k_{2}<\ldots$ be an infinite sequence of positive integers such that $h_{k_{0}} \equiv h_{k_{1}} \equiv \ldots(\bmod a)$. Notice that

$$
f\left(p_{k_{i}}-p_{k_{0}}\right)=f\left(p_{k_{i}}\right)-f\left(p_{k_{0}}\right)=h_{k_{i}}-h_{k_{0}} \equiv 0 \quad(\bmod a),
$$

so $p_{k_{i}}-p_{k_{0}} \in S_{a}$ for all $i=1,2, \ldots$. This provides infinitely many elements in $S_{a}$.
Hence, $S_{a}$ is an infinite set, and therefore $f(1)=a$ divides $f(n)$ for every $n$.

Comment. As an extension of the solution above, it can be proven that if $f$ is not bounded then $f(n)=a n$ with $a=f(1)$.

Take an arbitrary positive integer $n$; we will show that $f(n+1)=f(n)+a$. Then it follows by induction that $f(n)=a n$.

Take a peak $p$ such that $p>n+2$ and $h=f(p)>f(n)+2 a$. By (1) we have $f(p-1)=$ $f(p)-f(1)=h-a$ and $f(n+1)=f(p)-f(p-n-1)=h-f(p-n-1)$. From $h-a=f(p-1) \mid$ $f(n)+f(p-n-1)<f(n)+h<2(h-a)$ we get $f(n)+f(p-n-1)=h-a$. Then

$$
f(n+1)-f(n)=(h-f(p-n-1))-(h-a-f(p-n-1))=a .
$$

On the other hand, there exists a wide family of bounded functions satisfying the required properties. Here we present a few examples:

$$
f(n)=c ; \quad f(n)=\left\{\begin{array}{ll}
2 c & \text { if } n \text { is even } \\
c & \text { if } n \text { is odd } ;
\end{array} \quad f(n)= \begin{cases}2018 c & \text { if } n \leqslant 2018 \\
c & \text { if } n>2018\end{cases}\right.
$$

Solution 2. Let $d_{n}=\operatorname{gcd}(f(n), f(1))$. From $d_{n+1} \mid f(1)$ and $d_{n+1}|f(n+1)| f(n)+f(1)$, we can see that $d_{n+1} \mid f(n)$; then $d_{n+1} \mid \operatorname{gcd}(f(n), f(1))=d_{n}$. So the sequence $d_{1}, d_{2}, \ldots$ is nonincreasing in the sense that every element is a divisor of the previous elements. Let $d=\min \left(d_{1}, d_{2}, \ldots\right)=\operatorname{gcd}\left(d_{1} . d_{2}, \ldots\right)=\operatorname{gcd}(f(1), f(2), \ldots)$; we have to prove $d \geqslant 2$.

For the sake of contradiction, suppose that the statement is wrong, so $d=1$; that means there is some index $n_{0}$ such that $d_{n}=1$ for every $n \geqslant n_{0}$, i.e., $f(n)$ is coprime with $f(1)$.
Claim 1. If $2^{k} \geqslant n_{0}$ then $f\left(2^{k}\right) \leqslant 2^{k}$.
Proof. By the condition, $f(2 n) \mid 2 f(n)$; a trivial induction yields $f\left(2^{k}\right) \mid 2^{k} f(1)$. If $2^{k} \geqslant n_{0}$ then $f\left(2^{k}\right)$ is coprime with $f(1)$, so $f\left(2^{k}\right)$ is a divisor of $2^{k}$.
Claim 2. There is a constant $C$ such that $f(n)<n+C$ for every $n$.
Proof. Take the first power of 2 which is greater than or equal to $n_{0}$ : let $K=2^{k} \geqslant n_{0}$. By Claim 1, we have $f(K) \leqslant K$. Notice that $f(n+K) \mid f(n)+f(K)$ implies $f(n+K) \leqslant$ $f(n)+f(K) \leqslant f(n)+K$. If $n=t K+r$ for some $t \geqslant 0$ and $1 \leqslant r \leqslant K$, then we conclude
$f(n) \leqslant K+f(n-K) \leqslant 2 K+f(n-2 K) \leqslant \ldots \leqslant t K+f(r)<n+\max (f(1), f(2), \ldots, f(K))$, so the claim is true with $C=\max (f(1), \ldots, f(K))$.
Claim 3. If $a, b \in \mathbb{Z}_{>0}$ are coprime then $\operatorname{gcd}(f(a), f(b)) \mid f(1)$. In particular, if $a, b \geqslant n_{0}$ are coprime then $f(a)$ and $f(b)$ are coprime.
Proof. Let $d=\operatorname{gcd}(f(a), f(b))$. We can replicate Euclid's algorithm. Formally, apply induction on $a+b$. If $a=1$ or $b=1$ then we already have $d \mid f(1)$.

Without loss of generality, suppose $1<a<b$. Then $d \mid f(a)$ and $d|f(b)| f(a)+f(b-a)$, so $d \mid f(b-a)$. Therefore $d$ divides $\operatorname{gcd}(f(a), f(b-a))$ which is a divisor of $f(1)$ by the induction hypothesis.

Let $p_{1}<p_{2}<\ldots$ be the sequence of all prime numbers; for every $k$, let $q_{k}$ be the lowest power of $p_{k}$ with $q_{k} \geqslant n_{0}$. (Notice that there are only finitely many positive integers with $q_{k} \neq p_{k}$.)

Take a positive integer $N$, and consider the numbers

$$
f(1), f\left(q_{1}\right), f\left(q_{2}\right), \ldots, f\left(q_{N}\right)
$$

Here we have $N+1$ numbers, each being greater than 1 , and they are pairwise coprime by Claim 3. Therefore, they have at least $N+1$ different prime divisors in total, and their greatest prime divisor is at least $p_{N+1}$. Hence, $\max \left(f(1), f\left(q_{1}\right), \ldots, f\left(q_{N}\right)\right) \geqslant p_{N+1}$.

Choose $N$ such that $\max \left(q_{1}, \ldots, q_{N}\right)=p_{N}$ (this is achieved if $N$ is sufficiently large), and $p_{N+1}-p_{N}>C$ (that is possible, because there are arbitrarily long gaps between the primes). Then we establish a contradiction

$$
p_{N+1} \leqslant \max \left(f(1), f\left(q_{1}\right), \ldots, f\left(q_{N}\right)\right)<\max \left(1+C, q_{1}+C, \ldots, q_{N}+C\right)=p_{N}+C<p_{N+1}
$$

which proves the statement.

N7. Let $n \geqslant 2018$ be an integer, and let $a_{1}, a_{2}, \ldots, a_{n}, b_{1}, b_{2}, \ldots, b_{n}$ be pairwise distinct positive integers not exceeding $5 n$. Suppose that the sequence

$$
\begin{equation*}
\frac{a_{1}}{b_{1}}, \frac{a_{2}}{b_{2}}, \ldots, \frac{a_{n}}{b_{n}} \tag{1}
\end{equation*}
$$

forms an arithmetic progression. Prove that the terms of the sequence are equal.
(Thailand)
Solution. Suppose that (1) is an arithmetic progression with nonzero difference. Let the difference be $\Delta=\frac{c}{d}$, where $d>0$ and $c, d$ are coprime.

We will show that too many denominators $b_{i}$ should be divisible by $d$. To this end, for any $1 \leqslant i \leqslant n$ and any prime divisor $p$ of $d$, say that the index $i$ is $p$-wrong, if $v_{p}\left(b_{i}\right)<v_{p}(d) .\left(v_{p}(x)\right.$ stands for the exponent of $p$ in the prime factorisation of $x$.)
Claim 1. For any prime $p$, all $p$-wrong indices are congruent modulo $p$. In other words, the $p$-wrong indices (if they exist) are included in an arithmetic progression with difference $p$.
Proof. Let $\alpha=v_{p}(d)$. For the sake of contradiction, suppose that $i$ and $j$ are $p$-wrong indices (i.e., none of $b_{i}$ and $b_{j}$ is divisible by $p^{\alpha}$ ) such that $i \not \equiv j(\bmod p)$. Then the least common denominator of $\frac{a_{i}}{b_{i}}$ and $\frac{a_{j}}{b_{j}}$ is not divisible by $p^{\alpha}$. But this is impossible because in their difference, $(i-j) \Delta=\frac{(i-j) c}{d}$, the numerator is coprime to $p$, but $p^{\alpha}$ divides the denominator $d$.

Claim 2. $d$ has no prime divisors greater than 5 .
Proof. Suppose that $p \geqslant 7$ is a prime divisor of $d$. Among the indices $1,2, \ldots, n$, at most $\left\lceil\frac{n}{p}\right\rceil<\frac{n}{p}+1$ are $p$-wrong, so $p$ divides at least $\frac{p-1}{p} n-1$ of $b_{1}, \ldots, b_{n}$. Since these denominators are distinct,

$$
5 n \geqslant \max \left\{b_{i}: p \mid b_{i}\right\} \geqslant\left(\frac{p-1}{p} n-1\right) p=(p-1)(n-1)-1 \geqslant 6(n-1)-1>5 n
$$

a contradiction.
Claim 3. For every $0 \leqslant k \leqslant n-30$, among the denominators $b_{k+1}, b_{k+2}, \ldots, b_{k+30}$, at least $\varphi(30)=8$ are divisible by $d$.
Proof. By Claim 1, the 2-wrong, 3 -wrong and 5 -wrong indices can be covered by three arithmetic progressions with differences 2,3 and 5 . By a simple inclusion-exclusion, $(2-1) \cdot(3-1) \cdot(5-1)=8$ indices are not covered; by Claim 2, we have $d \mid b_{i}$ for every uncovered index $i$.

Claim 4. $|\Delta|<\frac{20}{n-2}$ and $d>\frac{n-2}{20}$.
Proof. From the sequence (1), remove all fractions with $b_{n}<\frac{n}{2}$, There remain at least $\frac{n}{2}$ fractions, and they cannot exceed $\frac{5 n}{n / 2}=10$. So we have at least $\frac{n}{2}$ elements of the arithmetic progression (1) in the interval $(0,10]$, hence the difference must be below $\frac{10}{n / 2-1}=\frac{20}{n-2}$.

The second inequality follows from $\frac{1}{d} \leqslant \frac{|c|}{d}=|\Delta|$.
Now we have everything to get the final contradiction. By Claim 3, we have $d \mid b_{i}$ for at least $\left\lfloor\frac{n}{30}\right\rfloor \cdot 8$ indices $i$. By Claim 4, we have $d \geqslant \frac{n-2}{20}$. Therefore,

$$
5 n \geqslant \max \left\{b_{i}: d \mid b_{i}\right\} \geqslant\left(\left\lfloor\frac{n}{30}\right\rfloor \cdot 8\right) \cdot d>\left(\frac{n}{30}-1\right) \cdot 8 \cdot \frac{n-2}{20}>5 n .
$$

Comment 1. It is possible that all terms in (1) are equal, for example with $a_{i}=2 i-1$ and $b_{i}=4 i-2$ we have $\frac{a_{i}}{b_{i}}=\frac{1}{2}$.

Comment 2. The bound $5 n$ in the statement is far from sharp; the solution above can be modified to work for $9 n$. For large $n$, the bound $5 n$ can be replaced by $n^{\frac{3}{2}-\varepsilon}$.

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## 60 ${ }^{\text {TH }}$ INTERNATIONAL MATHEMATICAL OLYMPIAD

July 11 ${ }^{\text {th }}$ - July $22^{\text {nd }}$, Bath, United Kingdom

# SHORTLISTED PROBLEMS WITH SOLUTIONS 



# Shortlisted Problems (with solutions) 

# The Shortlist has to be kept strictly confidential until the conclusion of the following International Mathematical Olympiad. 

 IMO General Regulations $\S 6.6$
## Contributing Countries

The Organising Committee and the Problem Selection Committee of IMO 2019 thank the following 58 countries for contributing 204 problem proposals:

Albania, Armenia, Australia, Austria, Belarus, Belgium, Brazil, Bulgaria, Canada, China, Croatia, Cuba, Cyprus, Czech Republic, Denmark, Ecuador, El Salvador, Estonia, Finland, France, Georgia, Germany, Greece, Hong Kong, Hungary, India, Indonesia, Iran, Ireland, Israel, Italy, Japan, Kazakhstan, Kosovo, Luxembourg, Mexico, Netherlands, New Zealand, Nicaragua, Nigeria, North Macedonia, Philippines, Poland, Russia, Serbia, Singapore, Slovakia, Slovenia, South Africa, South Korea, Sweden, Switzerland, Taiwan, Tanzania, Thailand, Ukraine, USA, Vietnam.

Problem Selection Committee


Tony Gardiner, Edward Crane, Alexander Betts, James Cranch, Joseph Myers (chair), James Aaronson, Andrew Carlotti, Géza Kós, Ilya I. Bogdanov, Jack Shotton

## Problems

## Algebra

A1. Let $\mathbb{Z}$ be the set of integers. Determine all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that, for all integers $a$ and $b$,

$$
f(2 a)+2 f(b)=f(f(a+b)) .
$$

(South Africa)
A2. Let $u_{1}, u_{2}, \ldots, u_{2019}$ be real numbers satisfying

$$
u_{1}+u_{2}+\cdots+u_{2019}=0 \quad \text { and } \quad u_{1}^{2}+u_{2}^{2}+\cdots+u_{2019}^{2}=1 .
$$

Let $a=\min \left(u_{1}, u_{2}, \ldots, u_{2019}\right)$ and $b=\max \left(u_{1}, u_{2}, \ldots, u_{2019}\right)$. Prove that

$$
a b \leqslant-\frac{1}{2019} .
$$

(Germany)
A3. Let $n \geqslant 3$ be a positive integer and let $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ be a strictly increasing sequence of $n$ positive real numbers with sum equal to 2 . Let $X$ be a subset of $\{1,2, \ldots, n\}$ such that the value of

$$
\left|1-\sum_{i \in X} a_{i}\right|
$$

is minimised. Prove that there exists a strictly increasing sequence of $n$ positive real numbers $\left(b_{1}, b_{2}, \ldots, b_{n}\right)$ with sum equal to 2 such that

$$
\sum_{i \in X} b_{i}=1 .
$$

(New Zealand)
A4. Let $n \geqslant 2$ be a positive integer and $a_{1}, a_{2}, \ldots, a_{n}$ be real numbers such that

$$
a_{1}+a_{2}+\cdots+a_{n}=0 .
$$

Define the set $A$ by

$$
A=\left\{(i, j)\left|1 \leqslant i<j \leqslant n,\left|a_{i}-a_{j}\right| \geqslant 1\right\} .\right.
$$

Prove that, if $A$ is not empty, then

$$
\sum_{(i, j) \in A} a_{i} a_{j}<0 .
$$

A5. Let $x_{1}, x_{2}, \ldots, x_{n}$ be different real numbers. Prove that

$$
\sum_{1 \leqslant i \leqslant n} \prod_{j \neq i} \frac{1-x_{i} x_{j}}{x_{i}-x_{j}}= \begin{cases}0, & \text { if } n \text { is even } \\ 1, & \text { if } n \text { is odd }\end{cases}
$$

A6. A polynomial $P(x, y, z)$ in three variables with real coefficients satisfies the identities

$$
P(x, y, z)=P(x, y, x y-z)=P(x, z x-y, z)=P(y z-x, y, z) .
$$

Prove that there exists a polynomial $F(t)$ in one variable such that

$$
P(x, y, z)=F\left(x^{2}+y^{2}+z^{2}-x y z\right)
$$

A7. Let $\mathbb{Z}$ be the set of integers. We consider functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ satisfying

$$
f(f(x+y)+y)=f(f(x)+y)
$$

for all integers $x$ and $y$. For such a function, we say that an integer $v$ is $f$-rare if the set

$$
X_{v}=\{x \in \mathbb{Z}: f(x)=v\}
$$

is finite and nonempty.
(a) Prove that there exists such a function $f$ for which there is an $f$-rare integer.
(b) Prove that no such function $f$ can have more than one $f$-rare integer.

## Combinatorics

C1. The infinite sequence $a_{0}, a_{1}, a_{2}, \ldots$ of (not necessarily different) integers has the following properties: $0 \leqslant a_{i} \leqslant i$ for all integers $i \geqslant 0$, and

$$
\binom{k}{a_{0}}+\binom{k}{a_{1}}+\cdots+\binom{k}{a_{k}}=2^{k}
$$

for all integers $k \geqslant 0$.
Prove that all integers $N \geqslant 0$ occur in the sequence (that is, for all $N \geqslant 0$, there exists $i \geqslant 0$ with $a_{i}=N$ ).
(Netherlands)
C2. You are given a set of $n$ blocks, each weighing at least 1 ; their total weight is $2 n$. Prove that for every real number $r$ with $0 \leqslant r \leqslant 2 n-2$ you can choose a subset of the blocks whose total weight is at least $r$ but at most $r+2$.
(Thailand)
C3. Let $n$ be a positive integer. Harry has $n$ coins lined up on his desk, each showing heads or tails. He repeatedly does the following operation: if there are $k$ coins showing heads and $k>0$, then he flips the $k^{\text {th }}$ coin over; otherwise he stops the process. (For example, the process starting with $T H T$ would be $T H T \rightarrow H H T \rightarrow H T T \rightarrow T T T$, which takes three steps.)

Letting $C$ denote the initial configuration (a sequence of $n H$ 's and $T$ 's), write $\ell(C)$ for the number of steps needed before all coins show $T$. Show that this number $\ell(C)$ is finite, and determine its average value over all $2^{n}$ possible initial configurations $C$.

C4. On a flat plane in Camelot, King Arthur builds a labyrinth $\mathfrak{L}$ consisting of $n$ walls, each of which is an infinite straight line. No two walls are parallel, and no three walls have a common point. Merlin then paints one side of each wall entirely red and the other side entirely blue.

At the intersection of two walls there are four corners: two diagonally opposite corners where a red side and a blue side meet, one corner where two red sides meet, and one corner where two blue sides meet. At each such intersection, there is a two-way door connecting the two diagonally opposite corners at which sides of different colours meet.

After Merlin paints the walls, Morgana then places some knights in the labyrinth. The knights can walk through doors, but cannot walk through walls.

Let $k(\mathfrak{L})$ be the largest number $k$ such that, no matter how Merlin paints the labyrinth $\mathfrak{L}$, Morgana can always place at least $k$ knights such that no two of them can ever meet. For each $n$, what are all possible values for $k(\mathfrak{L})$, where $\mathfrak{L}$ is a labyrinth with $n$ walls?
(Canada)
C5. On a certain social network, there are 2019 users, some pairs of which are friends, where friendship is a symmetric relation. Initially, there are 1010 people with 1009 friends each and 1009 people with 1010 friends each. However, the friendships are rather unstable, so events of the following kind may happen repeatedly, one at a time:

Let $A, B$, and $C$ be people such that $A$ is friends with both $B$ and $C$, but $B$ and $C$ are not friends; then $B$ and $C$ become friends, but $A$ is no longer friends with them.
Prove that, regardless of the initial friendships, there exists a sequence of such events after which each user is friends with at most one other user.

C6. Let $n>1$ be an integer. Suppose we are given $2 n$ points in a plane such that no three of them are collinear. The points are to be labelled $A_{1}, A_{2}, \ldots, A_{2 n}$ in some order. We then consider the $2 n$ angles $\angle A_{1} A_{2} A_{3}, \angle A_{2} A_{3} A_{4}, \ldots, \angle A_{2 n-2} A_{2 n-1} A_{2 n}, \angle A_{2 n-1} A_{2 n} A_{1}$, $\angle A_{2 n} A_{1} A_{2}$. We measure each angle in the way that gives the smallest positive value (i.e. between $0^{\circ}$ and $180^{\circ}$ ). Prove that there exists an ordering of the given points such that the resulting $2 n$ angles can be separated into two groups with the sum of one group of angles equal to the sum of the other group.

C7. There are 60 empty boxes $B_{1}, \ldots, B_{60}$ in a row on a table and an unlimited supply of pebbles. Given a positive integer $n$, Alice and Bob play the following game.

In the first round, Alice takes $n$ pebbles and distributes them into the 60 boxes as she wishes. Each subsequent round consists of two steps:
(a) Bob chooses an integer $k$ with $1 \leqslant k \leqslant 59$ and splits the boxes into the two groups $B_{1}, \ldots, B_{k}$ and $B_{k+1}, \ldots, B_{60}$.
(b) Alice picks one of these two groups, adds one pebble to each box in that group, and removes one pebble from each box in the other group.

Bob wins if, at the end of any round, some box contains no pebbles. Find the smallest $n$ such that Alice can prevent Bob from winning.
(Czech Republic)
C8. Alice has a map of Wonderland, a country consisting of $n \geqslant 2$ towns. For every pair of towns, there is a narrow road going from one town to the other. One day, all the roads are declared to be "one way" only. Alice has no information on the direction of the roads, but the King of Hearts has offered to help her. She is allowed to ask him a number of questions. For each question in turn, Alice chooses a pair of towns and the King of Hearts tells her the direction of the road connecting those two towns.

Alice wants to know whether there is at least one town in Wonderland with at most one outgoing road. Prove that she can always find out by asking at most $4 n$ questions.

Comment. This problem could be posed with an explicit statement about points being awarded for weaker bounds $c n$ for some $c>4$, in the style of IMO 2014 Problem 6.
(Thailand)
C9. For any two different real numbers $x$ and $y$, we define $D(x, y)$ to be the unique integer $d$ satisfying $2^{d} \leqslant|x-y|<2^{d+1}$. Given a set of reals $\mathcal{F}$, and an element $x \in \mathcal{F}$, we say that the scales of $x$ in $\mathcal{F}$ are the values of $D(x, y)$ for $y \in \mathcal{F}$ with $x \neq y$.

Let $k$ be a given positive integer. Suppose that each member $x$ of $\mathcal{F}$ has at most $k$ different scales in $\mathcal{F}$ (note that these scales may depend on $x$ ). What is the maximum possible size of $\mathcal{F}$ ?

## Geometry

G1. Let $A B C$ be a triangle. Circle $\Gamma$ passes through $A$, meets segments $A B$ and $A C$ again at points $D$ and $E$ respectively, and intersects segment $B C$ at $F$ and $G$ such that $F$ lies between $B$ and $G$. The tangent to circle $B D F$ at $F$ and the tangent to circle $C E G$ at $G$ meet at point $T$. Suppose that points $A$ and $T$ are distinct. Prove that line $A T$ is parallel to $B C$.
(Nigeria)
Q2. Let $A B C$ be an acute-angled triangle and let $D, E$, and $F$ be the feet of altitudes from $A, B$, and $C$ to sides $B C, C A$, and $A B$, respectively. Denote by $\omega_{B}$ and $\omega_{C}$ the incircles of triangles $B D F$ and $C D E$, and let these circles be tangent to segments $D F$ and $D E$ at $M$ and $N$, respectively. Let line $M N$ meet circles $\omega_{B}$ and $\omega_{C}$ again at $P \neq M$ and $Q \neq N$, respectively. Prove that $M P=N Q$.
(Vietnam)
G3. In triangle $A B C$, let $A_{1}$ and $B_{1}$ be two points on sides $B C$ and $A C$, and let $P$ and $Q$ be two points on segments $A A_{1}$ and $B B_{1}$, respectively, so that line $P Q$ is parallel to $A B$. On ray $P B_{1}$, beyond $B_{1}$, let $P_{1}$ be a point so that $\angle P P_{1} C=\angle B A C$. Similarly, on ray $Q A_{1}$, beyond $A_{1}$, let $Q_{1}$ be a point so that $\angle C Q_{1} Q=\angle C B A$. Show that points $P, Q, P_{1}$, and $Q_{1}$ are concyclic.
(Ukraine)
G4. Let $P$ be a point inside triangle $A B C$. Let $A P$ meet $B C$ at $A_{1}$, let $B P$ meet $C A$ at $B_{1}$, and let $C P$ meet $A B$ at $C_{1}$. Let $A_{2}$ be the point such that $A_{1}$ is the midpoint of $P A_{2}$, let $B_{2}$ be the point such that $B_{1}$ is the midpoint of $P B_{2}$, and let $C_{2}$ be the point such that $C_{1}$ is the midpoint of $P C_{2}$. Prove that points $A_{2}, B_{2}$, and $C_{2}$ cannot all lie strictly inside the circumcircle of triangle $A B C$.
(Australia)
G5. Let $A B C D E$ be a convex pentagon with $C D=D E$ and $\angle E D C \neq 2 \cdot \angle A D B$. Suppose that a point $P$ is located in the interior of the pentagon such that $A P=A E$ and $B P=B C$. Prove that $P$ lies on the diagonal $C E$ if and only if area $(B C D)+\operatorname{area}(A D E)=$ $\operatorname{area}(A B D)+\operatorname{area}(A B P)$.
(Hungary)
G6. Let $I$ be the incentre of acute-angled triangle $A B C$. Let the incircle meet $B C, C A$, and $A B$ at $D, E$, and $F$, respectively. Let line $E F$ intersect the circumcircle of the triangle at $P$ and $Q$, such that $F$ lies between $E$ and $P$. Prove that $\angle D P A+\angle A Q D=\angle Q I P$.
(Slovakia)


#### Abstract

G7. The incircle $\omega$ of acute-angled scalene triangle $A B C$ has centre $I$ and meets sides $B C$, $C A$, and $A B$ at $D, E$, and $F$, respectively. The line through $D$ perpendicular to $E F$ meets $\omega$ again at $R$. Line $A R$ meets $\omega$ again at $P$. The circumcircles of triangles $P C E$ and $P B F$ meet again at $Q \neq P$. Prove that lines $D I$ and $P Q$ meet on the external bisector of angle $B A C$.


(India)
G8. Let $\mathcal{L}$ be the set of all lines in the plane and let $f$ be a function that assigns to each line $\ell \in \mathcal{L}$ a point $f(\ell)$ on $\ell$. Suppose that for any point $X$, and for any three lines $\ell_{1}, \ell_{2}, \ell_{3}$ passing through $X$, the points $f\left(\ell_{1}\right), f\left(\ell_{2}\right), f\left(\ell_{3}\right)$ and $X$ lie on a circle.

Prove that there is a unique point $P$ such that $f(\ell)=P$ for any line $\ell$ passing through $P$.
(Australia)

## Number Theory

N1. Find all pairs ( $m, n$ ) of positive integers satisfying the equation

$$
\left(2^{n}-1\right)\left(2^{n}-2\right)\left(2^{n}-4\right) \cdots\left(2^{n}-2^{n-1}\right)=m!
$$

(El Salvador)
N2. Find all triples $(a, b, c)$ of positive integers such that $a^{3}+b^{3}+c^{3}=(a b c)^{2}$.
(Nigeria)
N3. We say that a set $S$ of integers is rootiful if, for any positive integer $n$ and any $a_{0}, a_{1}, \ldots, a_{n} \in S$, all integer roots of the polynomial $a_{0}+a_{1} x+\cdots+a_{n} x^{n}$ are also in $S$. Find all rootiful sets of integers that contain all numbers of the form $2^{a}-2^{b}$ for positive integers $a$ and $b$.
(Czech Republic)
N4. Let $\mathbb{Z}_{>0}$ be the set of positive integers. A positive integer constant $C$ is given. Find all functions $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ such that, for all positive integers $a$ and $b$ satisfying $a+b>C$,

$$
a+f(b) \mid a^{2}+b f(a)
$$

(Croatia)
N5. Let $a$ be a positive integer. We say that a positive integer $b$ is $a$-good if $\binom{a n}{b}-1$ is divisible by $a n+1$ for all positive integers $n$ with $a n \geqslant b$. Suppose $b$ is a positive integer such that $b$ is $a$-good, but $b+2$ is not $a$-good. Prove that $b+1$ is prime.
(Netherlands)
N6. Let $H=\left\{\lfloor i \sqrt{2}\rfloor: i \in \mathbb{Z}_{>0}\right\}=\{1,2,4,5,7, \ldots\}$, and let $n$ be a positive integer. Prove that there exists a constant $C$ such that, if $A \subset\{1,2, \ldots, n\}$ satisfies $|A| \geqslant C \sqrt{n}$, then there exist $a, b \in A$ such that $a-b \in H$. (Here $\mathbb{Z}_{>0}$ is the set of positive integers, and $\lfloor z\rfloor$ denotes the greatest integer less than or equal to $z$.)
(Brazil)

N7.
Prove that there is a constant $c>0$ and infinitely many positive integers $n$ with the following property: there are infinitely many positive integers that cannot be expressed as the sum of fewer than $c n \log (n)$ pairwise coprime $n^{\text {th }}$ powers.
(Canada)
N8.
Let $a$ and $b$ be two positive integers. Prove that the integer

$$
a^{2}+\left\lceil\frac{4 a^{2}}{b}\right\rceil
$$

is not a square. (Here $\lceil z\rceil$ denotes the least integer greater than or equal to $z$.)
(Russia)

## Solutions

## Algebra

A1. Let $\mathbb{Z}$ be the set of integers. Determine all functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ such that, for all integers $a$ and $b$,

$$
\begin{equation*}
f(2 a)+2 f(b)=f(f(a+b)) . \tag{1}
\end{equation*}
$$

(South Africa)
Answer: The solutions are $f(n)=0$ and $f(n)=2 n+K$ for any constant $K \in \mathbb{Z}$.
Common remarks. Most solutions to this problem first prove that $f$ must be linear, before determining all linear functions satisfying (1).

Solution 1. Substituting $a=0, b=n+1$ gives $f(f(n+1))=f(0)+2 f(n+1)$. Substituting $a=1, b=n$ gives $f(f(n+1))=f(2)+2 f(n)$.

In particular, $f(0)+2 f(n+1)=f(2)+2 f(n)$, and so $f(n+1)-f(n)=\frac{1}{2}(f(2)-f(0))$. Thus $f(n+1)-f(n)$ must be constant. Since $f$ is defined only on $\mathbb{Z}$, this tells us that $f$ must be a linear function; write $f(n)=M n+K$ for arbitrary constants $M$ and $K$, and we need only determine which choices of $M$ and $K$ work.

Now, (1) becomes

$$
2 M a+K+2(M b+K)=M(M(a+b)+K)+K
$$

which we may rearrange to form

$$
(M-2)(M(a+b)+K)=0 .
$$

Thus, either $M=2$, or $M(a+b)+K=0$ for all values of $a+b$. In particular, the only possible solutions are $f(n)=0$ and $f(n)=2 n+K$ for any constant $K \in \mathbb{Z}$, and these are easily seen to work.

Solution 2. Let $K=f(0)$.
First, put $a=0$ in (1); this gives

$$
\begin{equation*}
f(f(b))=2 f(b)+K \tag{2}
\end{equation*}
$$

for all $b \in \mathbb{Z}$.
Now put $b=0$ in (1); this gives

$$
f(2 a)+2 K=f(f(a))=2 f(a)+K,
$$

where the second equality follows from (2). Consequently,

$$
\begin{equation*}
f(2 a)=2 f(a)-K \tag{3}
\end{equation*}
$$

for all $a \in \mathbb{Z}$.
Substituting (2) and (3) into (1), we obtain

$$
\begin{aligned}
f(2 a)+2 f(b) & =f(f(a+b)) \\
2 f(a)-K+2 f(b) & =2 f(a+b)+K \\
f(a)+f(b) & =f(a+b)+K .
\end{aligned}
$$

Thus, if we set $g(n)=f(n)-K$ we see that $g$ satisfies the Cauchy equation $g(a+b)=$ $g(a)+g(b)$. The solution to the Cauchy equation over $\mathbb{Z}$ is well-known; indeed, it may be proven by an easy induction that $g(n)=M n$ for each $n \in \mathbb{Z}$, where $M=g(1)$ is a constant.

Therefore, $f(n)=M n+K$, and we may proceed as in Solution 1 .
Comment 1. Instead of deriving (3) by substituting $b=0$ into (1), we could instead have observed that the right hand side of (1) is symmetric in $a$ and $b$, and thus

$$
f(2 a)+2 f(b)=f(2 b)+2 f(a) .
$$

Thus, $f(2 a)-2 f(a)=f(2 b)-2 f(b)$ for any $a, b \in \mathbb{Z}$, and in particular $f(2 a)-2 f(a)$ is constant. Setting $a=0$ shows that this constant is equal to $-K$, and so we obtain (3).

Comment 2. Some solutions initially prove that $f(f(n))$ is linear (sometimes via proving that $f(f(n))-3 K$ satisfies the Cauchy equation). However, one can immediately prove that $f$ is linear by substituting something of the form $f(f(n))=M^{\prime} n+K^{\prime}$ into (2).

A2. Let $u_{1}, u_{2}, \ldots, u_{2019}$ be real numbers satisfying

$$
u_{1}+u_{2}+\cdots+u_{2019}=0 \quad \text { and } \quad u_{1}^{2}+u_{2}^{2}+\cdots+u_{2019}^{2}=1 .
$$

Let $a=\min \left(u_{1}, u_{2}, \ldots, u_{2019}\right)$ and $b=\max \left(u_{1}, u_{2}, \ldots, u_{2019}\right)$. Prove that

$$
a b \leqslant-\frac{1}{2019}
$$

(Germany)
Solution 1. Notice first that $b>0$ and $a<0$. Indeed, since $\sum_{i=1}^{2019} u_{i}^{2}=1$, the variables $u_{i}$ cannot be all zero, and, since $\sum_{i=1}^{2019} u_{i}=0$, the nonzero elements cannot be all positive or all negative.

Let $P=\left\{i: u_{i}>0\right\}$ and $N=\left\{i: u_{i} \leqslant 0\right\}$ be the indices of positive and nonpositive elements in the sequence, and let $p=|P|$ and $n=|N|$ be the sizes of these sets; then $p+n=2019$. By the condition $\sum_{i=1}^{2019} u_{i}=0$ we have $0=\sum_{i=1}^{2019} u_{i}=\sum_{i \in P} u_{i}-\sum_{i \in N}\left|u_{i}\right|$, so

$$
\begin{equation*}
\sum_{i \in P} u_{i}=\sum_{i \in N}\left|u_{i}\right| . \tag{1}
\end{equation*}
$$

After this preparation, estimate the sum of squares of the positive and nonpositive elements as follows:

$$
\begin{align*}
& \sum_{i \in P} u_{i}^{2} \leqslant \sum_{i \in P} b u_{i}=b \sum_{i \in P} u_{i}=b \sum_{i \in N}\left|u_{i}\right| \leqslant b \sum_{i \in N}|a|=-n a b ;  \tag{2}\\
& \sum_{i \in N} u_{i}^{2} \leqslant \sum_{i \in N}|a| \cdot\left|u_{i}\right|=|a| \sum_{i \in N}\left|u_{i}\right|=|a| \sum_{i \in P} u_{i} \leqslant|a| \sum_{i \in P} b=-p a b . \tag{3}
\end{align*}
$$

The sum of these estimates is

$$
1=\sum_{i=1}^{2019} u_{i}^{2}=\sum_{i \in P} u_{i}^{2}+\sum_{i \in N} u_{i}^{2} \leqslant-(p+n) a b=-2019 a b ;
$$

that proves $a b \leqslant \frac{-1}{2019}$.
Comment 1. After observing $\sum_{i \in P} u_{i}^{2} \leqslant b \sum_{i \in P} u_{i}$ and $\sum_{i \in N} u_{i}^{2} \leqslant|a| \sum_{i \in P}\left|u_{i}\right|$, instead of $(2,3)$ an alternative continuation is

$$
|a b| \geqslant \frac{\sum_{i \in P} u_{i}^{2}}{\sum_{i \in P} u_{i}} \cdot \frac{\sum_{i \in N} u_{i}^{2}}{\sum_{i \in N}\left|u_{i}\right|}=\frac{\sum_{i \in P} u_{i}^{2}}{\left(\sum_{i \in P} u_{i}\right)^{2}} \sum_{i \in N} u_{i}^{2} \geqslant \frac{1}{p} \sum_{i \in N} u_{i}^{2}
$$

(by the AM-QM or the Cauchy-Schwarz inequality) and similarly $|a b| \geqslant \frac{1}{n} \sum_{i \in P} u_{i}^{2}$.
Solution 2. As in the previous solution we conclude that $a<0$ and $b>0$.
For every index $i$, the number $u_{i}$ is a convex combination of $a$ and $b$, so

$$
u_{i}=x_{i} a+y_{i} b \quad \text { with some weights } 0 \leqslant x_{i}, y_{i} \leqslant 1, \text { with } x_{i}+y_{i}=1 \text {. }
$$

Let $X=\sum_{i=1}^{2019} x_{i}$ and $Y=\sum_{i=1}^{2019} y_{i}$. From $0=\sum_{i=1}^{2019} u_{i}=\sum_{i=1}^{2019}\left(x_{i} a+y_{i} b\right)=-|a| X+b Y$, we get

$$
\begin{equation*}
|a| X=b Y \tag{4}
\end{equation*}
$$

From $\sum_{i=1}^{2019}\left(x_{i}+y_{i}\right)=2019$ we have

$$
\begin{equation*}
X+Y=2019 \tag{5}
\end{equation*}
$$

The system of linear equations $(4,5)$ has a unique solution:

$$
X=\frac{2019 b}{|a|+b}, \quad Y=\frac{2019|a|}{|a|+b}
$$

Now apply the following estimate to every $u_{i}^{2}$ in their sum:

$$
u_{i}^{2}=x_{i}^{2} a^{2}+2 x_{i} y_{i} a b+y_{i}^{2} b^{2} \leqslant x_{i} a^{2}+y_{i} b^{2} ;
$$

we obtain that

$$
1=\sum_{i=1}^{2019} u_{i}^{2} \leqslant \sum_{i=1}^{2019}\left(x_{i} a^{2}+y_{i} b^{2}\right)=X a^{2}+Y b^{2}=\frac{2019 b}{|a|+b}|a|^{2}+\frac{2019|a|}{|a|+b} b^{2}=2019|a| b=-2019 a b .
$$

Hence, $a b \leqslant \frac{-1}{2019}$.
Comment 2. The idea behind Solution 2 is the following thought. Suppose we fix $a<0$ and $b>0$, fix $\sum u_{i}=0$ and vary the $u_{i}$ to achieve the maximum value of $\sum u_{i}^{2}$. Considering varying any two of the $u_{i}$ while preserving their sum: the maximum value of $\sum u_{i}^{2}$ is achieved when those two are as far apart as possible, so all but at most one of the $u_{i}$ are equal to $a$ or $b$. Considering a weighted version of the problem, we see the maximum (with fractional numbers of $u_{i}$ having each value) is achieved when $\frac{2019 b}{|a|+b}$ of them are $a$ and $\frac{2019|a|}{|a|+b}$ are $b$.

In fact, this happens in the solution: the number $u_{i}$ is replaced by $x_{i}$ copies of $a$ and $y_{i}$ copies of $b$.

A3. Let $n \geqslant 3$ be a positive integer and let $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ be a strictly increasing sequence of $n$ positive real numbers with sum equal to 2 . Let $X$ be a subset of $\{1,2, \ldots, n\}$ such that the value of

$$
\left|1-\sum_{i \in X} a_{i}\right|
$$

is minimised. Prove that there exists a strictly increasing sequence of $n$ positive real numbers $\left(b_{1}, b_{2}, \ldots, b_{n}\right)$ with sum equal to 2 such that

$$
\sum_{i \in X} b_{i}=1 .
$$

(New Zealand)
Common remarks. In all solutions, we say an index set $X$ is $\left(a_{i}\right)$-minimising if it has the property in the problem for the given sequence $\left(a_{i}\right)$. Write $X^{c}$ for the complement of $X$, and $[a, b]$ for the interval of integers $k$ such that $a \leqslant k \leqslant b$. Note that

$$
\left|1-\sum_{i \in X} a_{i}\right|=\left|1-\sum_{i \in X^{c}} a_{i}\right|,
$$

so we may exchange $X$ and $X^{c}$ where convenient. Let

$$
\Delta=\sum_{i \in X^{c}} a_{i}-\sum_{i \in X} a_{i}
$$

and note that $X$ is $\left(a_{i}\right)$-minimising if and only if it minimises $|\Delta|$, and that $\sum_{i \in X} a_{i}=1$ if and only if $\Delta=0$.

In some solutions, a scaling process is used. If we have a strictly increasing sequence of positive real numbers $c_{i}$ (typically obtained by perturbing the $a_{i}$ in some way) such that

$$
\sum_{i \in X} c_{i}=\sum_{i \in X^{c}} c_{i}
$$

then we may put $b_{i}=2 c_{i} / \sum_{j=1}^{n} c_{j}$. So it suffices to construct such a sequence without needing its sum to be 2 .

The solutions below show various possible approaches to the problem. Solutions 1 and 2 perturb a few of the $a_{i}$ to form the $b_{i}$ (with scaling in the case of Solution 1, without scaling in the case of Solution 2). Solutions 3 and 4 look at properties of the index set $X$. Solution 3 then perturbs many of the $a_{i}$ to form the $b_{i}$, together with scaling. Rather than using such perturbations, Solution 4 constructs a sequence $\left(b_{i}\right)$ directly from the set $X$ with the required properties. Solution 4 can be used to give a complete description of sets $X$ that are $\left(a_{i}\right)$-minimising for some $\left(a_{i}\right)$.

Solution 1. Without loss of generality, assume $\sum_{i \in X} a_{i} \leqslant 1$, and we may assume strict inequality as otherwise $b_{i}=a_{i}$ works. Also, $X$ clearly cannot be empty.

If $n \in X$, add $\Delta$ to $a_{n}$, producing a sequence of $c_{i}$ with $\sum_{i \in X} c_{i}=\sum_{i \in X^{c}} c_{i}$, and then scale as described above to make the sum equal to 2 . Otherwise, there is some $k$ with $k \in X$ and $k+1 \in X^{c}$. Let $\delta=a_{k+1}-a_{k}$.

- If $\delta>\Delta$, add $\Delta$ to $a_{k}$ and then scale.
- If $\delta<\Delta$, then considering $X \cup\{k+1\} \backslash\{k\}$ contradicts $X$ being $\left(a_{i}\right)$-minimising.
- If $\delta=\Delta$, choose any $j \neq k, k+1$ (possible since $n \geqslant 3$ ), and any $\epsilon$ less than the least of $a_{1}$ and all the differences $a_{i+1}-a_{i}$. If $j \in X$ then add $\Delta-\epsilon$ to $a_{k}$ and $\epsilon$ to $a_{j}$, then scale; otherwise, add $\Delta$ to $a_{k}$ and $\epsilon / 2$ to $a_{k+1}$, and subtract $\epsilon / 2$ from $a_{j}$, then scale.

Solution 2. This is similar to Solution 1, but without scaling. As in that solution, without loss of generality, assume $\sum_{i \in X} a_{i}<1$.

Suppose there exists $1 \leqslant j \leqslant n-1$ such that $j \in X$ but $j+1 \in X^{c}$. Then $a_{j+1}-a_{j} \geqslant \Delta$, because otherwise considering $X \cup\{j+1\} \backslash\{j\}$ contradicts $X$ being $\left(a_{i}\right)$-minimising.

If $a_{j+1}-a_{j}>\Delta$, put

$$
b_{i}= \begin{cases}a_{j}+\Delta / 2, & \text { if } i=j \\ a_{j+1}-\Delta / 2, & \text { if } i=j+1 \\ a_{i}, & \text { otherwise }\end{cases}
$$

If $a_{j+1}-a_{j}=\Delta$, choose any $\epsilon$ less than the least of $\Delta / 2, a_{1}$ and all the differences $a_{i+1}-a_{i}$. If $|X| \geqslant 2$, choose $k \in X$ with $k \neq j$, and put

$$
b_{i}= \begin{cases}a_{j}+\Delta / 2-\epsilon, & \text { if } i=j \\ a_{j+1}-\Delta / 2, & \text { if } i=j+1 \\ a_{k}+\epsilon, & \text { if } i=k \\ a_{i}, & \text { otherwise }\end{cases}
$$

Otherwise, $\left|X^{c}\right| \geqslant 2$, so choose $k \in X^{c}$ with $k \neq j+1$, and put

$$
b_{i}= \begin{cases}a_{j}+\Delta / 2, & \text { if } i=j \\ a_{j+1}-\Delta / 2+\epsilon, & \text { if } i=j+1 \\ a_{k}-\epsilon, & \text { if } i=k ; \\ a_{i}, & \text { otherwise }\end{cases}
$$

If there is no $1 \leqslant j \leqslant n$ such that $j \in X$ but $j+1 \in X^{c}$, there must be some $1<k \leqslant n$ such that $X=[k, n]$ (certainly $X$ cannot be empty). We must have $a_{1}>\Delta$, as otherwise considering $X \cup\{1\}$ contradicts $X$ being $\left(a_{i}\right)$-minimising. Now put

$$
b_{i}= \begin{cases}a_{1}-\Delta / 2, & \text { if } i=1 \\ a_{n}+\Delta / 2, & \text { if } i=n \\ a_{i}, & \text { otherwise }\end{cases}
$$

Solution 3. Without loss of generality, assume $\sum_{i \in X} a_{i} \leqslant 1$, so $\Delta \geqslant 0$. If $\Delta=0$ we can take $b_{i}=a_{i}$, so now assume that $\Delta>0$.

Suppose that there is some $k \leqslant n$ such that $|X \cap[k, n]|>\left|X^{c} \cap[k, n]\right|$. If we choose the largest such $k$ then $|X \cap[k, n]|-\left|X^{c} \cap[k, n]\right|=1$. We can now find the required sequence $\left(b_{i}\right)$ by starting with $c_{i}=a_{i}$ for $i<k$ and $c_{i}=a_{i}+\Delta$ for $i \geqslant k$, and then scaling as described above.

If no such $k$ exists, we will derive a contradiction. For each $i \in X$ we can choose $i<j_{i} \leqslant n$ in such a way that $j_{i} \in X^{c}$ and all the $j_{i}$ are different. (For instance, note that necessarily $n \in X^{c}$ and now just work downwards; each time an $i \in X$ is considered, let $j_{i}$ be the least element of $X^{c}$ greater than $i$ and not yet used.) Let $Y$ be the (possibly empty) subset of $[1, n]$ consisting of those elements in $X^{c}$ that are also not one of the $j_{i}$. In any case

$$
\Delta=\sum_{i \in X}\left(a_{j_{i}}-a_{i}\right)+\sum_{j \in Y} a_{j}
$$

where each term in the sums is positive. Since $n \geqslant 3$ the total number of terms above is at least two. Take a least such term and its corresponding index $i$ and consider the set $Z$ which we form from $X$ by removing $i$ and adding $j_{i}$ (if it is a term of the first type) or just by adding $j$ if it is a term of the second type. The corresponding expression of $\Delta$ for $Z$ has the sign of its least term changed, meaning that the sum is still nonnegative but strictly less than $\Delta$, which contradicts $X$ being $\left(a_{i}\right)$-minimising.

Solution 4. This uses some similar ideas to Solution 3, but describes properties of the index sets $X$ that are sufficient to describe a corresponding sequence $\left(b_{i}\right)$ that is not derived from $\left(a_{i}\right)$.

Note that, for two subsets $X, Y$ of $[1, n]$, the following are equivalent:

- $|X \cap[i, n]| \leqslant|Y \cap[i, n]|$ for all $1 \leqslant i \leqslant n$;
- $Y$ is at least as large as $X$, and for all $1 \leqslant j \leqslant|Y|$, the $j^{\text {th }}$ largest element of $Y$ is at least as big as the $j^{\text {th }}$ largest element of $X$;
- there is an injective function $f: X \rightarrow Y$ such that $f(i) \geqslant i$ for all $i \in X$.

If these equivalent conditions are satisfied, we write $X \leq Y$. We write $X<Y$ if $X \leq Y$ and $X \neq Y$.

Note that if $X<Y$, then $\sum_{i \in X} a_{i}<\sum_{i \in Y} a_{i}$ (the second description above makes this clear).
We claim first that, if $n \geqslant 3$ and $X<X^{c}$, then there exists $Y$ with $X<Y<X^{c}$. Indeed, as $|X| \leqslant\left|X^{c}\right|$, we have $\left|X^{c}\right| \geqslant 2$. Define $Y$ to consist of the largest element of $X^{c}$, together with all but the largest element of $X$; it is clear both that $Y$ is distinct from $X$ and $X^{c}$, and that $X \leq Y \leq X^{c}$, which is what we need.

But, in this situation, we have

$$
\sum_{i \in X} a_{i}<\sum_{i \in Y} a_{i}<\sum_{i \in X^{c}} a_{i} \quad \text { and } \quad 1-\sum_{i \in X} a_{i}=-\left(1-\sum_{i \in X^{c}} a_{i}\right)
$$

so $\left|1-\sum_{i \in Y} a_{i}\right|<\left|1-\sum_{i \in X} a_{i}\right|$.
Hence if $X$ is $\left(a_{i}\right)$-minimising, we do not have $X<X^{c}$, and similarly we do not have $X^{c}<X$.

Considering the first description above, this immediately implies the following Claim.
Claim. There exist $1 \leqslant k, \ell \leqslant n$ such that $|X \cap[k, n]|>\frac{n-k+1}{2}$ and $|X \cap[\ell, n]|<\frac{n-\ell+1}{2}$.
We now construct our sequence $\left(b_{i}\right)$ using this claim. Let $k$ and $\ell$ be the greatest values satisfying the claim, and without loss of generality suppose $k=n$ and $\ell<n$ (otherwise replace $X$ by its complement). As $\ell$ is maximal, $n-\ell$ is even and $|X \cap[\ell, n]|=\frac{n-\ell}{2}$. For sufficiently small positive $\epsilon$, we take

$$
b_{i}=i \epsilon+ \begin{cases}0, & \text { if } i<\ell \\ \delta, & \text { if } \ell \leqslant i \leqslant n-1 \\ \gamma, & \text { if } i=n\end{cases}
$$

Let $M=\sum_{i \in X} i$. So we require

$$
M \epsilon+\left(\frac{n-\ell}{2}-1\right) \delta+\gamma=1
$$

and

$$
\frac{n(n+1)}{2} \epsilon+(n-\ell) \delta+\gamma=2
$$

These give

$$
\gamma=2 \delta+\left(\frac{n(n+1)}{2}-2 M\right) \epsilon
$$

and for sufficiently small positive $\epsilon$, solving for $\gamma$ and $\delta$ gives $0<\delta<\gamma$ (since $\epsilon=0$ gives $\delta=1 /\left(\frac{n-\ell}{2}+1\right)$ and $\left.\gamma=2 \delta\right)$, so the sequence is strictly increasing and has positive values.

Comment. This solution also shows that the claim gives a complete description of sets $X$ that are $\left(a_{i}\right)$-minimising for some $\left(a_{i}\right)$.

Another approach to proving the claim is as follows. We prove the existence of $\ell$ with the claimed property; the existence of $k$ follows by considering the complement of $X$.

Suppose, for a contradiction, that for all $1 \leqslant \ell \leqslant n$ we have $|X \cap[\ell, n]| \geqslant\left\lceil\frac{n-\ell+1}{2}\right\rceil$. If we ever have strict inequality, consider the set $Y=\{n, n-2, n-4, \ldots\}$. This set may be obtained from $X$ by possibly removing some elements and reducing the values of others. (To see this, consider the largest $k \in X \backslash Y$, if any; remove it, and replace it by the greatest $j \in X^{c}$ with $j<k$, if any. Such steps preserve the given inequality, and are possible until we reach the set $Y$.) So if we had strict inequality, and so $X \neq Y$, we have

$$
\sum_{i \in X} a_{i}>\sum_{i \in Y} a_{i}>1,
$$

contradicting $X$ being $\left(a_{i}\right)$-minimising. Otherwise, we always have equality, meaning that $X=Y$. But now consider $Z=Y \cup\{n-1\} \backslash\{n\}$. Since $n \geqslant 3$, we have

$$
\sum_{i \in Y} a_{i}>\sum_{i \in Z} a_{i}>\sum_{i \in Y^{c}} a_{i}=2-\sum_{i \in Y} a_{i},
$$

and so $Z$ contradicts $X$ being $\left(a_{i}\right)$-minimising.

A4. Let $n \geqslant 2$ be a positive integer and $a_{1}, a_{2}, \ldots, a_{n}$ be real numbers such that

$$
a_{1}+a_{2}+\cdots+a_{n}=0 .
$$

Define the set $A$ by

$$
A=\left\{(i, j)\left|1 \leqslant i<j \leqslant n,\left|a_{i}-a_{j}\right| \geqslant 1\right\} .\right.
$$

Prove that, if $A$ is not empty, then

$$
\sum_{(i, j) \in A} a_{i} a_{j}<0 .
$$

(China)
Solution 1. Define sets $B$ and $C$ by

$$
\begin{aligned}
& B=\left\{(i, j)\left|1 \leqslant i, j \leqslant n,\left|a_{i}-a_{j}\right| \geqslant 1\right\},\right. \\
& C=\left\{(i, j)\left|1 \leqslant i, j \leqslant n,\left|a_{i}-a_{j}\right|<1\right\} .\right.
\end{aligned}
$$

We have

$$
\begin{aligned}
\sum_{(i, j) \in A} a_{i} a_{j} & =\frac{1}{2} \sum_{(i, j) \in B} a_{i} a_{j} \\
\sum_{(i, j) \in B} a_{i} a_{j} & =\sum_{1 \leqslant i, j \leqslant n} a_{i} a_{j}-\sum_{(i, j) \notin B} a_{i} a_{j}=0-\sum_{(i, j) \in C} a_{i} a_{j} .
\end{aligned}
$$

So it suffices to show that if $A$ (and hence $B$ ) are nonempty, then

$$
\sum_{(i, j) \in C} a_{i} a_{j}>0 .
$$

Partition the indices into sets $P, Q, R$, and $S$ such that

$$
\begin{aligned}
P & =\left\{i \mid a_{i} \leqslant-1\right\} \\
Q & =\left\{i \mid-1<a_{i} \leqslant 0\right\}
\end{aligned}
$$

$$
R=\left\{i \mid 0<a_{i}<1\right\}
$$

$$
S=\left\{i \mid 1 \leqslant a_{i}\right\} .
$$

Then

$$
\sum_{(i, j) \in C} a_{i} a_{j} \geqslant \sum_{i \in P \cup S} a_{i}^{2}+\sum_{i, j \in Q \cup R} a_{i} a_{j}=\sum_{i \in P \cup S} a_{i}^{2}+\left(\sum_{i \in Q \cup R} a_{i}\right)^{2} \geqslant 0 .
$$

The first inequality holds because all of the positive terms in the RHS are also in the LHS, and all of the negative terms in the LHS are also in the RHS. The first inequality attains equality only if both sides have the same negative terms, which implies $\left|a_{i}-a_{j}\right|<1$ whenever $i, j \in Q \cup R$; the second inequality attains equality only if $P=S=\varnothing$. But then we would have $A=\varnothing$. So $A$ nonempty implies that the inequality holds strictly, as required.

Solution 2. Consider $P, Q, R, S$ as in Solution 1, set

$$
p=\sum_{i \in P} a_{i}, \quad q=\sum_{i \in Q} a_{i}, \quad r=\sum_{i \in R} a_{i}, \quad s=\sum_{i \in S} a_{i},
$$

and let

$$
t_{+}=\sum_{(i, j) \in A, a_{i} a_{j} \geqslant 0} a_{i} a_{j}, \quad t_{-}=\sum_{(i, j) \in A, a_{i} a_{j} \leqslant 0} a_{i} a_{j} .
$$

We know that $p+q+r+s=0$, and we need to prove that $t_{+}+t_{-}<0$.
Notice that $t_{+} \leqslant p^{2} / 2+p q+r s+s^{2} / 2$ (with equality only if $p=s=0$ ), and $t_{-} \leqslant p r+p s+q s$ (with equality only if there do not exist $i \in Q$ and $j \in R$ with $a_{j}-a_{i}>1$ ). Therefore,

$$
t_{+}+t_{-} \leqslant \frac{p^{2}+s^{2}}{2}+p q+r s+p r+p s+q s=\frac{(p+q+r+s)^{2}}{2}-\frac{(q+r)^{2}}{2}=-\frac{(q+r)^{2}}{2} \leqslant 0
$$

If $A$ is not empty and $p=s=0$, then there must exist $i \in Q, j \in R$ with $\left|a_{i}-a_{j}\right|>1$, and hence the earlier equality conditions cannot both occur.

Comment. The RHS of the original inequality cannot be replaced with any constant $c<0$ (independent of $n$ ). Indeed, take

$$
a_{1}=-\frac{n}{n+2}, a_{2}=\cdots=a_{n-1}=\frac{1}{n+2}, a_{n}=\frac{2}{n+2} .
$$

Then $\sum_{(i, j) \in A} a_{i} a_{j}=-\frac{2 n}{(n+2)^{2}}$, which converges to zero as $n \rightarrow \infty$.

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A5. Let $x_{1}, x_{2}, \ldots, x_{n}$ be different real numbers. Prove that

$$
\sum_{1 \leqslant i \leqslant n} \prod_{j \neq i} \frac{1-x_{i} x_{j}}{x_{i}-x_{j}}= \begin{cases}0, & \text { if } n \text { is even } \\ 1, & \text { if } n \text { is odd }\end{cases}
$$

(Kazakhstan)
Common remarks. Let $G\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ be the function of the $n$ variables $x_{1}, x_{2}, \ldots, x_{n}$ on the LHS of the required identity.

Solution 1 (Lagrange interpolation). Since both sides of the identity are rational functions, it suffices to prove it when all $x_{i} \notin\{ \pm 1\}$. Define

$$
f(t)=\prod_{i=1}^{n}\left(1-x_{i} t\right)
$$

and note that

$$
f\left(x_{i}\right)=\left(1-x_{i}^{2}\right) \prod_{j \neq i} 1-x_{i} x_{j} .
$$

Using the nodes $+1,-1, x_{1}, \ldots, x_{n}$, the Lagrange interpolation formula gives us the following expression for $f$ :

$$
\sum_{i=1}^{n} f\left(x_{i}\right) \frac{(x-1)(x+1)}{\left(x_{i}-1\right)\left(x_{i}+1\right)} \prod_{j \neq i} \frac{x-x_{j}}{x_{i}-x_{j}}+f(1) \frac{x+1}{1+1} \prod_{1 \leqslant i \leqslant n} \frac{x-x_{i}}{1-x_{i}}+f(-1) \frac{x-1}{-1-1} \prod_{1 \leqslant i \leqslant n} \frac{x-x_{i}}{1-x_{i}}
$$

The coefficient of $t^{n+1}$ in $f(t)$ is 0 , since $f$ has degree $n$. The coefficient of $t^{n+1}$ in the above expression of $f$ is

$$
\begin{aligned}
0 & =\sum_{1 \leqslant i \leqslant n} \frac{f\left(x_{i}\right)}{\prod_{j \neq i}\left(x_{i}-x_{j}\right) \cdot\left(x_{i}-1\right)\left(x_{i}+1\right)}+\frac{f(1)}{\prod_{1 \leqslant j \leqslant n}\left(1-x_{j}\right) \cdot(1+1)}+\frac{f(-1)}{\prod_{1 \leqslant j \leqslant n}\left(-1-x_{j}\right) \cdot(-1-1)} \\
& =-G\left(x_{1}, \ldots, x_{n}\right)+\frac{1}{2}+\frac{(-1)^{n+1}}{2} .
\end{aligned}
$$

Comment. The main difficulty is to think of including the two extra nodes $\pm 1$ and evaluating the coefficient $t^{n+1}$ in $f$ when $n+1$ is higher than the degree of $f$.

It is possible to solve the problem using Lagrange interpolation on the nodes $x_{1}, \ldots, x_{n}$, but the definition of the polynomial being interpolated should depend on the parity of $n$. For $n$ even, consider the polynomial

$$
P(x)=\prod_{i}\left(1-x x_{i}\right)-\prod_{i}\left(x-x_{i}\right) .
$$

Lagrange interpolation shows that $G$ is the coefficient of $x^{n-1}$ in the polynomial $P(x) /\left(1-x^{2}\right)$, i.e. 0 . For $n$ odd, consider the polynomial

$$
P(x)=\prod_{i}\left(1-x x_{i}\right)-x \prod_{i}\left(x-x_{i}\right) .
$$

Now $G$ is the coefficient of $x^{n-1}$ in $P(x) /\left(1-x^{2}\right)$, which is 1 .

Solution 2 (using symmetries). Observe that $G$ is symmetric in the variables $x_{1}, \ldots, x_{n}$. Define $V=\prod_{i<j}\left(x_{j}-x_{i}\right)$ and let $F=G \cdot V$, which is a polynomial in $x_{1}, \ldots, x_{n}$. Since $V$ is alternating, $F$ is also alternating (meaning that, if we exchange any two variables, then $F$ changes sign). Every alternating polynomial in $n$ variables $x_{1}, \ldots, x_{n}$ vanishes when any two variables $x_{i}, x_{j}(i \neq j)$ are equal, and is therefore divisible by $x_{i}-x_{j}$ for each pair $i \neq j$. Since these linear factors are pairwise coprime, $V$ divides $F$ exactly as a polynomial. Thus $G$ is in fact a symmetric polynomial in $x_{1}, \ldots, x_{n}$.

Now observe that if all $x_{i}$ are nonzero and we set $y_{i}=1 / x_{i}$ for $i=1, \ldots, n$, then we have

$$
\frac{1-y_{i} y_{j}}{y_{i}-y_{j}}=\frac{1-x_{i} x_{j}}{x_{i}-x_{j}}
$$

so that

$$
G\left(\frac{1}{x_{1}}, \ldots, \frac{1}{x_{n}}\right)=G\left(x_{1}, \ldots, x_{n}\right)
$$

By continuity this is an identity of rational functions. Since $G$ is a polynomial, it implies that $G$ is constant. (If $G$ were not constant, we could choose a point $\left(c_{1}, \ldots, c_{n}\right)$ with all $c_{i} \neq 0$, such that $G\left(c_{1}, \ldots, c_{n}\right) \neq G(0, \ldots, 0)$; then $g(x):=G\left(c_{1} x, \ldots, c_{n} x\right)$ would be a nonconstant polynomial in the variable $x$, so $|g(x)| \rightarrow \infty$ as $x \rightarrow \infty$, hence $\left|G\left(\frac{y}{c_{1}}, \ldots, \frac{y}{c_{n}}\right)\right| \rightarrow \infty$ as $y \rightarrow 0$, which is impossible since $G$ is a polynomial.)

We may identify the constant by substituting $x_{i}=\zeta^{i}$, where $\zeta$ is a primitive $n^{\text {th }}$ root of unity in $\mathbb{C}$. In the $i^{\text {th }}$ term in the sum in the original expression we have a factor $1-\zeta^{i} \zeta^{n-i}=0$, unless $i=n$ or $2 i=n$. In the case where $n$ is odd, the only exceptional term is $i=n$, which gives the value $\prod_{j \neq n} \frac{1-\zeta^{j}}{1-\zeta^{j}}=1$. When $n$ is even, we also have the term $\prod_{j \neq \frac{n}{2}} \frac{1+\zeta^{j}}{-1-\zeta^{j}}=(-1)^{n-1}=-1$, so the sum is 0 .

Comment. If we write out an explicit expression for $F$,

$$
F=\sum_{1 \leqslant i \leqslant n}(-1)^{n-i} \prod_{\substack{j<k \\ j, k \neq i}}\left(x_{k}-x_{j}\right) \prod_{j \neq i}\left(1-x_{i} x_{j}\right)
$$

then to prove directly that $F$ vanishes when $x_{i}=x_{j}$ for some $i \neq j$, but no other pair of variables coincide, we have to check carefully that the two nonzero terms in this sum cancel.

A different and slightly less convenient way to identify the constant is to substitute $x_{i}=1+\epsilon \zeta^{i}$, and throw away terms that are $O(\epsilon)$ as $\epsilon \rightarrow 0$.

Solution 3 (breaking symmetry). Consider $G$ as a rational function in $x_{n}$ with coefficients that are rational functions in the other variables. We can write

$$
G\left(x_{1}, \ldots, x_{n}\right)=\frac{P\left(x_{n}\right)}{\prod_{j \neq n}\left(x_{n}-x_{j}\right)}
$$

where $P\left(x_{n}\right)$ is a polynomial in $x_{n}$ whose coefficients are rational functions in the other variables. We then have

$$
P\left(x_{n}\right)=\left(\prod_{j \neq n}\left(1-x_{n} x_{j}\right)\right)+\sum_{1 \leqslant i \leqslant n-1}\left(x_{i} x_{n}-1\right)\left(\prod_{j \neq i, n}\left(x_{n}-x_{j}\right)\right)\left(\prod_{j \neq i, n} \frac{1-x_{i} x_{j}}{x_{i}-x_{j}}\right) .
$$

For any $k \neq n$, substituting $x_{n}=x_{k}$ (which is valid when manipulating the numerator $P\left(x_{n}\right)$
on its own), we have (noting that $x_{n}-x_{j}$ vanishes when $j=k$ )

$$
\begin{aligned}
P\left(x_{k}\right) & =\left(\prod_{j \neq n}\left(1-x_{k} x_{j}\right)\right)+\sum_{1 \leqslant i \leqslant n-1}\left(x_{i} x_{k}-1\right)\left(\prod_{j \neq i, n}\left(x_{k}-x_{j}\right)\right)\left(\prod_{j \neq i, n} \frac{1-x_{i} x_{j}}{x_{i}-x_{j}}\right) \\
& =\left(\prod_{j \neq n}\left(1-x_{k} x_{j}\right)\right)+\left(x_{k}^{2}-1\right)\left(\prod_{j \neq k, n}\left(x_{k}-x_{j}\right)\right)\left(\prod_{j \neq k, n} \frac{1-x_{k} x_{j}}{x_{k}-x_{j}}\right) \\
& =\left(\prod_{j \neq n}\left(1-x_{k} x_{j}\right)\right)+\left(x_{k}^{2}-1\right)\left(\prod_{j \neq k, n}\left(1-x_{k} x_{j}\right)\right) \\
& =0 .
\end{aligned}
$$

Note that $P$ is a polynomial in $x_{n}$ of degree $n-1$. For any choice of distinct real numbers $x_{1}, \ldots, x_{n-1}, P$ has those real numbers as its roots, and the denominator has the same degree and the same roots. This shows that $G$ is constant in $x_{n}$, for any fixed choice of distinct $x_{1}, \ldots, x_{n-1}$. Now, $G$ is symmetric in all $n$ variables, so it must be also be constant in each of the other variables. $G$ is therefore a constant that depends only on $n$. The constant may be identified as in the previous solution.

Comment. There is also a solution in which we recognise the expression for $F$ in the comment after Solution 2 as the final column expansion of a certain matrix obtained by modifying the final column of the Vandermonde matrix. The task is then to show that the matrix can be modified by column operations either to make the final column identically zero (in the case where $n$ even) or to recover the Vandermonde matrix (in the case where $n$ odd). The polynomial $P /\left(1-x^{2}\right)$ is helpful for this task, where $P$ is the parity-dependent polynomial defined in the comment after Solution 1.

A6. A polynomial $P(x, y, z)$ in three variables with real coefficients satisfies the identities

$$
\begin{equation*}
P(x, y, z)=P(x, y, x y-z)=P(x, z x-y, z)=P(y z-x, y, z) . \tag{*}
\end{equation*}
$$

Prove that there exists a polynomial $F(t)$ in one variable such that

$$
P(x, y, z)=F\left(x^{2}+y^{2}+z^{2}-x y z\right) .
$$

(Russia)
Common remarks. The polynomial $x^{2}+y^{2}+z^{2}-x y z$ satisfies the condition (*), so every polynomial of the form $F\left(x^{2}+y^{2}+z^{2}-x y z\right)$ does satisfy (*). We will use without comment the fact that two polynomials have the same coefficients if and only if they are equal as functions.

Solution 1. In the first two steps, we deal with any polynomial $P(x, y, z)$ satisfying $P(x, y, z)=$ $P(x, y, x y-z)$. Call such a polynomial weakly symmetric, and call a polynomial satisfying the full conditions in the problem symmetric.

Step 1. We start with the description of weakly symmetric polynomials. We claim that they are exactly the polynomials in $x, y$, and $z(x y-z)$. Clearly, all such polynomials are weakly symmetric. For the converse statement, consider $P_{1}(x, y, z):=P\left(x, y, z+\frac{1}{2} x y\right)$, which satisfies $P_{1}(x, y, z)=P_{1}(x, y,-z)$ and is therefore a polynomial in $x, y$, and $z^{2}$. This means that $P$ is a polynomial in $x, y$, and $\left(z-\frac{1}{2} x y\right)^{2}=-z(x y-z)+\frac{1}{4} x^{2} y^{2}$, and therefore a polynomial in $x, y$, and $z(x y-z)$.

Step 2. Suppose that $P$ is weakly symmetric. Consider the monomials in $P(x, y, z)$ of highest total degree. Our aim is to show that in each such monomial $\mu x^{a} y^{b} z^{c}$ we have $a, b \geqslant c$. Consider the expansion

$$
\begin{equation*}
P(x, y, z)=\sum_{i, j, k} \mu_{i j k} x^{i} y^{j}(z(x y-z))^{k} \tag{1.1}
\end{equation*}
$$

The maximal total degree of a summand in (1.1) is $m=\max _{i, j, k: \mu_{i j k} \neq 0}(i+j+3 k)$. Now, for any $i, j, k$ satisfying $i+j+3 k=m$ the summand $\mu_{i, j, k} x^{i} y^{j}(z(x y-z))^{k}$ has leading term of the form $\mu x^{i+k} y^{j+k} z^{k}$. No other nonzero summand in (1.1) may have a term of this form in its expansion, hence this term does not cancel in the whole sum. Therefore, $\operatorname{deg} P=m$, and the leading component of $P$ is exactly

$$
\sum_{i+j+3 k=m} \mu_{i, j, k} x^{i+k} y^{j+k} z^{k}
$$

and each summand in this sum satisfies the condition claimed above.
Step 3. We now prove the problem statement by induction on $m=\operatorname{deg} P$. For $m=0$ the claim is trivial. Consider now a symmetric polynomial $P$ with $\operatorname{deg} P>0$. By Step 2, each of its monomials $\mu x^{a} y^{b} z^{c}$ of the highest total degree satisfies $a, b \geqslant c$. Applying other weak symmetries, we obtain $a, c \geqslant b$ and $b, c \geqslant a$; therefore, $P$ has a unique leading monomial of the form $\mu(x y z)^{c}$. The polynomial $P_{0}(x, y, z)=P(x, y, z)-\mu\left(x y z-x^{2}-y^{2}-z^{2}\right)^{c}$ has smaller total degree. Since $P_{0}$ is symmetric, it is representable as a polynomial function of $x y z-x^{2}-y^{2}-z^{2}$. Then $P$ is also of this form, completing the inductive step.

Comment. We could alternatively carry out Step 1 by an induction on $n=\operatorname{deg}_{z} P$, in a manner similar to Step 3. If $n=0$, the statement holds. Assume that $n>0$ and check the leading component of $P$ with respect to $z$ :

$$
P(x, y, z)=Q_{n}(x, y) z^{n}+R(x, y, z),
$$

where $\operatorname{deg}_{z} R<n$. After the change $z \mapsto x y-z$, the leading component becomes $Q_{n}(x, y)(-z)^{n}$; on the other hand, it should remain the same. Hence $n$ is even. Now consider the polynomial

$$
P_{0}(x, y, z)=P(x, y, z)-Q_{n}(x, y) \cdot(z(z-x y))^{n / 2}
$$

It is also weakly symmetric, and $\operatorname{deg}_{z} P_{0}<n$. By the inductive hypothesis, it has the form $P_{0}(x, y, z)=$ $S(x, y, z(z-x y))$. Hence the polynomial

$$
P(x, y, z)=S(x, y, z(x y-z))+Q_{n}(x, y)(z(z-x y))^{n / 2}
$$

also has this form. This completes the inductive step.
Solution 2. We will rely on the well-known identity

$$
\begin{equation*}
\cos ^{2} u+\cos ^{2} v+\cos ^{2} w-2 \cos u \cos v \cos w-1=0 \quad \text { whenever } u+v+w=0 \tag{2.1}
\end{equation*}
$$

Claim 1. The polynomial $P(x, y, z)$ is constant on the surface

$$
\mathfrak{S}=\{(2 \cos u, 2 \cos v, 2 \cos w): u+v+w=0\}
$$

Proof. Notice that for $x=2 \cos u, y=2 \cos v, z=2 \cos w$, the Vieta jumps $x \mapsto y z-x$, $y \mapsto z x-y, z \mapsto x y-z$ in $(*)$ replace $(u, v, w)$ by $(v-w,-v, w),(u, w-u,-w)$ and $(-u, v, u-v)$, respectively. For example, for the first type of jump we have

$$
y z-x=4 \cos v \cos w-2 \cos u=2 \cos (v+w)+2 \cos (v-w)-2 \cos u=2 \cos (v-w) .
$$

Define $G(u, v, w)=P(2 \cos u, 2 \cos v, 2 \cos w)$. For $u+v+w=0$, the jumps give

$$
\begin{aligned}
G(u, v, w) & =G(v-w,-v, w)=G(w-v,-v,(v-w)-(-v))=G(-u-2 v,-v, 2 v-w) \\
& =G(u+2 v, v, w-2 v) .
\end{aligned}
$$

By induction,

$$
\begin{equation*}
G(u, v, w)=G(u+2 k v, v, w-2 k v) \quad(k \in \mathbb{Z}) . \tag{2.2}
\end{equation*}
$$

Similarly,

$$
\begin{equation*}
G(u, v, w)=G(u, v-2 \ell u, w+2 \ell u) \quad(\ell \in \mathbb{Z}) \tag{2.3}
\end{equation*}
$$

And, of course, we have

$$
\begin{equation*}
G(u, v, w)=G(u+2 p \pi, v+2 q \pi, w-2(p+q) \pi) \quad(p, q \in \mathbb{Z}) \tag{2.4}
\end{equation*}
$$

Take two nonzero real numbers $u, v$ such that $u, v$ and $\pi$ are linearly independent over $\mathbb{Q}$. By combining (2.2-2.4), we can see that $G$ is constant on a dense subset of the plane $u+v+w=0$. By continuity, $G$ is constant on the entire plane and therefore $P$ is constant on $\mathfrak{S}$.
Claim 2. The polynomial $T(x, y, z)=x^{2}+y^{2}+z^{2}-x y z-4$ divides $P(x, y, z)-P(2,2,2)$.
Proof. By dividing $P$ by $T$ with remainders, there exist some polynomials $R(x, y, z), A(y, z)$ and $B(y, z)$ such that

$$
\begin{equation*}
P(x, y, z)-P(2,2,2)=T(x, y, z) \cdot R(x, y, z)+A(y, z) x+B(y, z) \tag{2.5}
\end{equation*}
$$

On the surface $\mathfrak{S}$ the LHS of (2.5) is zero by Claim 1 (since $(2,2,2) \in \mathfrak{S}$ ) and $T=0$ by (2.1). Hence, $A(y, z) x+B(y, z)$ vanishes on $\mathfrak{S}$.

Notice that for every $y=2 \cos v$ and $z=2 \cos w$ with $\frac{\pi}{3}<v, w<\frac{2 \pi}{3}$, there are two distinct values of $x$ such that $(x, y, z) \in \mathfrak{S}$, namely $x_{1}=2 \cos (v+w)$ (which is negative), and $x_{2}=2 \cos (v-w)$ (which is positive). This can happen only if $A(y, z)=B(y, z)=0$. Hence, $A(y, z)=B(y, z)=0$ for $|y|<1,|z|<1$. The polynomials $A$ and $B$ vanish on an open set, so $A$ and $B$ are both the zero polynomial.

The quotient $(P(x, y, z)-P(2,2,2)) / T(x, y, z)$ is a polynomial of lower degree than $P$ and it also satisfies (*). The problem statement can now be proven by induction on the degree of $P$.

Comment. In the proof of (2.2) and (2.3) we used two consecutive Vieta jumps; in fact from (*) we used only $P(x, y, x y-z)=P(x, z x-y, z)=P(y z-x, y, z)$.

Solution 3 (using algebraic geometry, just for interest). Let $Q=x^{2}+y^{2}+z^{2}-x y z$ and let $t \in \mathbb{C}$. Checking where $Q-t, \frac{\partial Q}{\partial x}, \frac{\partial Q}{\partial y}$ and $\frac{\partial Q}{\partial z}$ vanish simultaneously, we find that the surface $Q=t$ is smooth except for the cases $t=0$, when the only singular point is $(0,0,0)$, and $t=4$, when the four points $( \pm 2, \pm 2, \pm 2)$ that satisfy $x y z=8$ are the only singular points. The singular points are the fixed points of the group $\Gamma$ of polynomial automorphisms of $\mathbb{C}^{3}$ generated by the three Vieta involutions

$$
\iota_{1}:(x, y, z) \mapsto(x, y, x y-z), \quad \iota_{2}:(x, y, z) \mapsto(x, x z-y, z), \quad \iota_{3}:(x, y, z) \mapsto(y z-x, y, z) .
$$

$\Gamma$ acts on each surface $\mathcal{V}_{t}: Q-t=0$. If $Q-t$ were reducible then the surface $Q=t$ would contain a curve of singular points. Therefore $Q-t$ is irreducible in $\mathbb{C}[x, y, z]$. (One can also prove algebraically that $Q-t$ is irreducible, for example by checking that its discriminant as a quadratic polynomial in $x$ is not a square in $\mathbb{C}[y, z]$, and likewise for the other two variables.) In the following solution we will only use the algebraic surface $\mathcal{V}_{0}$.

Let $U$ be the $\Gamma$-orbit of $(3,3,3)$. Consider $\iota_{3} \circ \iota_{2}$, which leaves $z$ invariant. For each fixed value of $z, \iota_{3} \circ \iota_{2}$ acts linearly on $(x, y)$ by the matrix

$$
M_{z}:=\left(\begin{array}{cc}
z^{2}-1 & -z \\
z & -1
\end{array}\right) .
$$

The reverse composition $\iota_{2} \circ \iota_{3}$ acts by $M_{z}^{-1}=M_{z}^{\text {adj }}$. Note det $M_{z}=1$ and $\operatorname{tr} M_{z}=z^{2}-2$. When $z$ does not lie in the real interval $[-2,2]$, the eigenvalues of $M_{z}$ do not have absolute value 1, so every orbit of the group generated by $M_{z}$ on $\mathbb{C}^{2} \backslash\{(0,0)\}$ is unbounded. For example, fixing $z=3$ we find $\left(3 F_{2 k+1}, 3 F_{2 k-1}, 3\right) \in U$ for every $k \in \mathbb{Z}$, where $\left(F_{n}\right)_{n \in \mathbb{Z}}$ is the Fibonacci sequence with $F_{0}=0, F_{1}=1$.

Now we may start at any point $\left(3 F_{2 k+1}, 3 F_{2 k-1}, 3\right)$ and iteratively apply $\iota_{1} \circ \iota_{2}$ to generate another infinite sequence of distinct points of $U$, Zariski dense in the hyperbola cut out of $\mathcal{V}_{0}$ by the plane $x-3 F_{2 k+1}=0$. (The plane $x=a$ cuts out an irreducible conic when $a \notin\{-2,0,2\}$.) Thus the Zariski closure $\bar{U}$ of $U$ contains infinitely many distinct algebraic curves in $\mathcal{V}_{0}$. Since $\mathcal{V}_{0}$ is an irreducible surface this implies that $\bar{U}=\mathcal{V}_{0}$.

For any polynomial $P$ satisfying (*), we have $P-P(3,3,3)=0$ at each point of $U$. Since $\bar{U}=\mathcal{V}_{0}, P-P(3,3,3)$ vanishes on $\mathcal{V}_{0}$. Then Hilbert's Nullstellensatz and the irreducibility of $Q$ imply that $P-P(3,3,3)$ is divisible by $Q$. Now $(P-P(3,3,3)) / Q$ is a polynomial also satisfying (*), so we may complete the proof by an induction on the total degree, as in the other solutions.

Comment. We remark that Solution 2 used a trigonometric parametrisation of a real component of $\mathcal{V}_{4}$; in contrast $\mathcal{V}_{0}$ is birationally equivalent to the projective space $\mathbb{P}^{2}$ under the maps

$$
(x, y, z) \rightarrow(x: y: z), \quad(a: b: c) \rightarrow\left(\frac{a^{2}+b^{2}+c^{2}}{b c}, \frac{a^{2}+b^{2}+c^{2}}{a c}, \frac{a^{2}+b^{2}+c^{2}}{a b}\right) .
$$

The set $U$ in Solution 3 is contained in $\mathbb{Z}^{3}$ so it is nowhere dense in $\mathcal{V}_{0}$ in the classical topology.
Comment (background to the problem). A triple $(a, b, c) \in \mathbb{Z}^{3}$ is called a Markov triple if $a^{2}+b^{2}+c^{2}=3 a b c$, and an integer that occurs as a coordinate of some Markov triple is called a Markov number. (The spelling Markoff is also frequent.) Markov triples arose in A. Markov's work in the 1870s on the reduction theory of indefinite binary quadratic forms. For every Markov triple,
$(3 a, 3 b, 3 c)$ lies on $Q=0$. It is well known that all nonzero Markov triples can be generated from $(1,1,1)$ by sequences of Vieta involutions, which are the substitutions described in equation $(*)$ in the problem statement. There has been recent work by number theorists about the properties of Markov numbers (see for example Jean Bourgain, Alex Gamburd and Peter Sarnak, Markoff triples and strong approximation, Comptes Rendus Math. 345, no. 2, 131-135 (2016), arXiv:1505.06411). Each Markov number occurs in infinitely many triples, but a famous old open problem is the unicity conjecture, which asserts that each Markov number occurs in only one Markov triple (up to permutations and sign changes) as the largest coordinate in absolute value in that triple. It is a standard fact in the modern literature on Markov numbers that the Markov triples are Zariski dense in the Markov surface. Proving this is the main work of Solution 3. Algebraic geometry is definitely off-syllabus for the IMO, and one still has to work a bit to prove the Zariski density. On the other hand the approaches of Solutions 1 and 2 are elementary and only use tools expected to be known by IMO contestants. Therefore we do not think that the existence of a solution using algebraic geometry necessarily makes this problem unsuitable for the IMO.

A7. Let $\mathbb{Z}$ be the set of integers. We consider functions $f: \mathbb{Z} \rightarrow \mathbb{Z}$ satisfying

$$
f(f(x+y)+y)=f(f(x)+y)
$$

for all integers $x$ and $y$. For such a function, we say that an integer $v$ is $f$-rare if the set

$$
X_{v}=\{x \in \mathbb{Z}: f(x)=v\}
$$

is finite and nonempty.
(a) Prove that there exists such a function $f$ for which there is an $f$-rare integer.
(b) Prove that no such function $f$ can have more than one $f$-rare integer.
(Netherlands)
Solution 1. a) Let $f$ be the function where $f(0)=0$ and $f(x)$ is the largest power of 2 dividing $2 x$ for $x \neq 0$. The integer 0 is evidently $f$-rare, so it remains to verify the functional equation.

Since $f(2 x)=2 f(x)$ for all $x$, it suffices to verify the functional equation when at least one of $x$ and $y$ is odd (the case $x=y=0$ being trivial). If $y$ is odd, then we have

$$
f(f(x+y)+y)=2=f(f(x)+y)
$$

since all the values attained by $f$ are even. If, on the other hand, $x$ is odd and $y$ is even, then we already have

$$
f(x+y)=2=f(x)
$$

from which the functional equation follows immediately.
b) An easy inductive argument (substituting $x+k y$ for $x$ ) shows that

$$
\begin{equation*}
f(f(x+k y)+y)=f(f(x)+y) \tag{*}
\end{equation*}
$$

for all integers $x, y$ and $k$. If $v$ is an $f$-rare integer and $a$ is the least element of $X_{v}$, then by substituting $y=a-f(x)$ in the above, we see that

$$
f(x+k \cdot(a-f(x)))-f(x)+a \in X_{v}
$$

for all integers $x$ and $k$, so that in particular

$$
f(x+k \cdot(a-f(x))) \geqslant f(x)
$$

for all integers $x$ and $k$, by assumption on $a$. This says that on the (possibly degenerate) arithmetic progression through $x$ with common difference $a-f(x)$, the function $f$ attains its minimal value at $x$.

Repeating the same argument with $a$ replaced by the greatest element $b$ of $X_{v}$ shows that

$$
f(x+k \cdot(b-f(x)) \leqslant f(x)
$$

for all integers $x$ and $k$. Combined with the above inequality, we therefore have

$$
f(x+k \cdot(a-f(x)) \cdot(b-f(x)))=f(x)
$$

for all integers $x$ and $k$.
Thus if $f(x) \neq a, b$, then the set $X_{f(x)}$ contains a nondegenerate arithmetic progression, so is infinite. So the only possible $f$-rare integers are $a$ and $b$.

In particular, the $f$-rare integer $v$ we started with must be one of $a$ or $b$, so that $f(v)=$ $f(a)=f(b)=v$. This means that there cannot be any other $f$-rare integers $v^{\prime}$, as they would on the one hand have to be either $a$ or $b$, and on the other would have to satisfy $f\left(v^{\prime}\right)=v^{\prime}$. Thus $v$ is the unique $f$-rare integer.

Comment 1. If $f$ is a solution to the functional equation, then so too is any conjugate of $f$ by a translation, i.e. any function $x \mapsto f(x+n)-n$ for an integer $n$. Thus in proving part (b), one is free to consider only functions $f$ for which 0 is $f$-rare, as in the following solution.

Solution 2, part (b) only. Suppose $v$ is $f$-rare, and let $a$ and $b$ be the least and greatest elements of $X_{v}$, respectively. Substituting $x=v$ and $y=a-v$ into the equation shows that

$$
f(v)-v+a \in X_{v}
$$

and in particular $f(v) \geqslant v$. Repeating the same argument with $x=v$ and $y=b-v$ shows that $f(v) \leqslant v$, and hence $f(v)=v$.

Suppose now that $v^{\prime}$ is a second $f$-rare integer. We may assume that $v=0$ (see Comment 1 ). We've seen that $f\left(v^{\prime}\right)=v^{\prime}$; we claim that in fact $f\left(k v^{\prime}\right)=v^{\prime}$ for all positive integers $k$. This gives a contradiction unless $v^{\prime}=v=0$.

This claim is proved by induction on $k$. Supposing it to be true for $k$, we substitute $y=k v^{\prime}$ and $x=0$ into the functional equation to yield

$$
f\left((k+1) v^{\prime}\right)=f\left(f(0)+k v^{\prime}\right)=f\left(k v^{\prime}\right)=v^{\prime}
$$

using that $f(0)=0$. This completes the induction, and hence the proof.
Comment 2. There are many functions $f$ satisfying the functional equation for which there is an $f$-rare integer. For instance, one may generalise the construction in part (a) of Solution 1 by taking a sequence $1=a_{0}, a_{1}, a_{2}, \ldots$ of positive integers with each $a_{i}$ a proper divisor of $a_{i+1}$ and choosing arbitrary functions $f_{i}:\left(\mathbb{Z} / a_{i} \mathbb{Z}\right) \backslash\{0\} \rightarrow a_{i} \mathbb{Z} \backslash\{0\}$ from the nonzero residue classes modulo $a_{i}$ to the nonzero multiples of $a_{i}$. One then defines a function $f: \mathbb{Z} \rightarrow \mathbb{Z}$ by

$$
f(x):= \begin{cases}f_{i+1}\left(x \bmod a_{i+1}\right), & \text { if } a_{i} \mid x \text { but } a_{i+1} \nmid x ; \\ 0, & \text { if } x=0\end{cases}
$$

If one writes $v(x)$ for the largest $i$ such that $a_{i} \mid x$ (with $v(0)=\infty$ ), then it is easy to verify the functional equation for $f$ separately in the two cases $v(y)>v(x)$ and $v(x) \geqslant v(y)$. Hence this $f$ satisfies the functional equation and 0 is an $f$-rare integer.

Comment 3. In fact, if $v$ is an $f$-rare integer for an $f$ satisfying the functional equation, then its fibre $X_{v}=\{v\}$ must be a singleton. We may assume without loss of generality that $v=0$. We've already seen in Solution 1 that 0 is either the greatest or least element of $X_{0}$; replacing $f$ with the function $x \mapsto-f(-x)$ if necessary, we may assume that 0 is the least element of $X_{0}$. We write $b$ for the largest element of $X_{0}$, supposing for contradiction that $b>0$, and write $N=(2 b)!$.

It now follows from $(*)$ that we have

$$
f(f(N b)+b)=f(f(0)+b)=f(b)=0
$$

from which we see that $f(N b)+b \in X_{0} \subseteq[0, b]$. It follows that $f(N b) \in[-b, 0)$, since by construction $N b \notin X_{v}$. Now it follows that $(f(N b)-0) \cdot(f(N b)-b)$ is a divisor of $N$, so from $(\dagger)$ we see that $f(N b)=f(0)=0$. This yields the desired contradiction.

## Combinatorics

C1. The infinite sequence $a_{0}, a_{1}, a_{2}, \ldots$ of (not necessarily different) integers has the following properties: $0 \leqslant a_{i} \leqslant i$ for all integers $i \geqslant 0$, and

$$
\binom{k}{a_{0}}+\binom{k}{a_{1}}+\cdots+\binom{k}{a_{k}}=2^{k}
$$

for all integers $k \geqslant 0$.
Prove that all integers $N \geqslant 0$ occur in the sequence (that is, for all $N \geqslant 0$, there exists $i \geqslant 0$ with $\left.a_{i}=N\right)$.
(Netherlands)
Solution. We prove by induction on $k$ that every initial segment of the sequence, $a_{0}, a_{1}, \ldots, a_{k}$, consists of the following elements (counted with multiplicity, and not necessarily in order), for some $\ell \geqslant 0$ with $2 \ell \leqslant k+1$ :

$$
0,1, \ldots, \ell-1, \quad 0,1, \ldots, k-\ell
$$

For $k=0$ we have $a_{0}=0$, which is of this form. Now suppose that for $k=m$ the elements $a_{0}, a_{1}, \ldots, a_{m}$ are $0,0,1,1,2,2, \ldots, \ell-1, \ell-1, \ell, \ell+1, \ldots, m-\ell-1, m-\ell$ for some $\ell$ with $0 \leqslant 2 \ell \leqslant m+1$. It is given that

$$
\binom{m+1}{a_{0}}+\binom{m+1}{a_{1}}+\cdots+\binom{m+1}{a_{m}}+\binom{m+1}{a_{m+1}}=2^{m+1}
$$

which becomes

$$
\begin{aligned}
\left(\binom{m+1}{0}+\binom{m+1}{1}\right. & \left.+\cdots+\binom{m+1}{\ell-1}\right) \\
& +\left(\binom{m+1}{0}+\binom{m+1}{1}+\cdots+\binom{m+1}{m-\ell}\right)+\binom{m+1}{a_{m+1}}=2^{m+1}
\end{aligned}
$$

or, using $\binom{m+1}{i}=\binom{m+1}{m+1-i}$, that

$$
\begin{aligned}
\left(\binom{m+1}{0}+\binom{m+1}{1}\right. & \left.+\cdots+\binom{m+1}{\ell-1}\right) \\
& +\left(\binom{m+1}{m+1}+\binom{m+1}{m}+\cdots+\binom{m+1}{\ell+1}\right)+\binom{m+1}{a_{m+1}}=2^{m+1}
\end{aligned}
$$

On the other hand, it is well known that

$$
\binom{m+1}{0}+\binom{m+1}{1}+\cdots+\binom{m+1}{m+1}=2^{m+1}
$$

and so, by subtracting, we get

$$
\binom{m+1}{a_{m+1}}=\binom{m+1}{\ell} .
$$

From this, using the fact that the binomial coefficients $\binom{m+1}{i}$ are increasing for $i \leqslant \frac{m+1}{2}$ and decreasing for $i \geqslant \frac{m+1}{2}$, we conclude that either $a_{m+1}=\ell$ or $a_{m+1}=m+1-\ell$. In either case, $a_{0}, a_{1}, \ldots, a_{m+1}$ is again of the claimed form, which concludes the induction.

As a result of this description, any integer $N \geqslant 0$ appears as a term of the sequence $a_{i}$ for some $0 \leqslant i \leqslant 2 N$.

C2. You are given a set of $n$ blocks, each weighing at least 1 ; their total weight is $2 n$. Prove that for every real number $r$ with $0 \leqslant r \leqslant 2 n-2$ you can choose a subset of the blocks whose total weight is at least $r$ but at most $r+2$.
(Thailand)
Solution 1. We prove the following more general statement by induction on $n$.
Claim. Suppose that you have $n$ blocks, each of weight at least 1 , and of total weight $s \leqslant 2 n$. Then for every $r$ with $-2 \leqslant r \leqslant s$, you can choose some of the blocks whose total weight is at least $r$ but at most $r+2$.
Proof. The base case $n=1$ is trivial. To prove the inductive step, let $x$ be the largest block weight. Clearly, $x \geqslant s / n$, so $s-x \leqslant \frac{n-1}{n} s \leqslant 2(n-1)$. Hence, if we exclude a block of weight $x$, we can apply the inductive hypothesis to show the claim holds (for this smaller set) for any $-2 \leqslant r \leqslant s-x$. Adding the excluded block to each of those combinations, we see that the claim also holds when $x-2 \leqslant r \leqslant s$. So if $x-2 \leqslant s-x$, then we have covered the whole interval $[-2, s]$. But each block weight is at least 1 , so we have $x-2 \leqslant(s-(n-1))-2=s-(2 n-(n-1)) \leqslant s-(s-(n-1)) \leqslant s-x$, as desired.

Comment. Instead of inducting on sets of blocks with total weight $s \leqslant 2 n$, we could instead prove the result only for $s=2 n$. We would then need to modify the inductive step to scale up the block weights before applying the induction hypothesis.

Solution 2. Let $x_{1}, \ldots, x_{n}$ be the weights of the blocks in weakly increasing order. Consider the set $S$ of sums of the form $\sum_{j \in J} x_{j}$ for a subset $J \subseteq\{1,2, \ldots, n\}$. We want to prove that the mesh of $S$ - i.e. the largest distance between two adjacent elements - is at most 2.

For $0 \leqslant k \leqslant n$, let $S_{k}$ denote the set of sums of the form $\sum_{i \in J} x_{i}$ for a subset $J \subseteq\{1,2, \ldots, k\}$. We will show by induction on $k$ that the mesh of $S_{k}$ is at most 2 .

The base case $k=0$ is trivial (as $S_{0}=\{0\}$ ). For $k>0$ we have

$$
S_{k}=S_{k-1} \cup\left(x_{k}+S_{k-1}\right)
$$

(where $\left(x_{k}+S_{k-1}\right)$ denotes $\left\{x_{k}+s: s \in S_{k-1}\right\}$ ), so it suffices to prove that $x_{k} \leqslant \sum_{j<k} x_{j}+2$. But if this were not the case, we would have $x_{l}>\sum_{j<k} x_{j}+2 \geqslant k+1$ for all $l \geqslant k$, and hence

$$
2 n=\sum_{j=1}^{n} x_{j}>(n+1-k)(k+1)+k-1 .
$$

This rearranges to $n>k(n+1-k)$, which is false for $1 \leqslant k \leqslant n$, giving the desired contradiction.

## C3. Let $n$ be a positive integer. Harry has $n$ coins lined up on his desk, each showing

 heads or tails. He repeatedly does the following operation: if there are $k$ coins showing heads and $k>0$, then he flips the $k^{\text {th }}$ coin over; otherwise he stops the process. (For example, the process starting with THT would be THT $\rightarrow H H T \rightarrow H T T \rightarrow T T T$, which takes three steps.)Letting $C$ denote the initial configuration (a sequence of $n H$ 's and $T$ 's), write $\ell(C)$ for the number of steps needed before all coins show $T$. Show that this number $\ell(C)$ is finite, and determine its average value over all $2^{n}$ possible initial configurations $C$.

Answer: The average is $\frac{1}{4} n(n+1)$.
Common remarks. Throughout all these solutions, we let $E(n)$ denote the desired average value.

Solution 1. We represent the problem using a directed graph $G_{n}$ whose vertices are the length- $n$ strings of $H$ 's and $T$ 's. The graph features an edge from each string to its successor (except for $T T \cdots T T$, which has no successor). We will also write $\bar{H}=T$ and $\bar{T}=H$.

The graph $G_{0}$ consists of a single vertex: the empty string. The main claim is that $G_{n}$ can be described explicitly in terms of $G_{n-1}$ :

- We take two copies, $X$ and $Y$, of $G_{n-1}$.
- In $X$, we take each string of $n-1$ coins and just append a $T$ to it. In symbols, we replace $s_{1} \cdots s_{n-1}$ with $s_{1} \cdots s_{n-1} T$.
- In $Y$, we take each string of $n-1$ coins, flip every coin, reverse the order, and append an $H$ to it. In symbols, we replace $s_{1} \cdots s_{n-1}$ with $\bar{s}_{n-1} \bar{s}_{n-2} \cdots \bar{s}_{1} H$.
- Finally, we add one new edge from $Y$ to $X$, namely $H H \cdots H H H \rightarrow H H \cdots H H T$.

We depict $G_{4}$ below, in a way which indicates this recursive construction:


We prove the claim inductively. Firstly, $X$ is correct as a subgraph of $G_{n}$, as the operation on coins is unchanged by an extra $T$ at the end: if $s_{1} \cdots s_{n-1}$ is sent to $t_{1} \cdots t_{n-1}$, then $s_{1} \cdots s_{n-1} T$ is sent to $t_{1} \cdots t_{n-1} T$.

Next, $Y$ is also correct as a subgraph of $G_{n}$, as if $s_{1} \cdots s_{n-1}$ has $k$ occurrences of $H$, then $\bar{s}_{n-1} \cdots \bar{s}_{1} H$ has $(n-1-k)+1=n-k$ occurrences of $H$, and thus (provided that $k>0$ ), if $s_{1} \cdots s_{n-1}$ is sent to $t_{1} \cdots t_{n-1}$, then $\bar{s}_{n-1} \cdots \bar{s}_{1} H$ is sent to $\bar{t}_{n-1} \cdots \bar{t}_{1} H$.

Finally, the one edge from $Y$ to $X$ is correct, as the operation does send $H H \cdots H H H$ to $H H \cdots H H T$.

To finish, note that the sequences in $X$ take an average of $E(n-1)$ steps to terminate, whereas the sequences in $Y$ take an average of $E(n-1)$ steps to reach $H H \cdots H$ and then an additional $n$ steps to terminate. Therefore, we have

$$
E(n)=\frac{1}{2}(E(n-1)+(E(n-1)+n))=E(n-1)+\frac{n}{2} .
$$

We have $E(0)=0$ from our description of $G_{0}$. Thus, by induction, we have $E(n)=\frac{1}{2}(1+\cdots+$ $n)=\frac{1}{4} n(n+1)$, which in particular is finite.

Solution 2. We consider what happens with configurations depending on the coins they start and end with.

- If a configuration starts with $H$, the last $n-1$ coins follow the given rules, as if they were all the coins, until they are all $T$, then the first coin is turned over.
- If a configuration ends with $T$, the last coin will never be turned over, and the first $n-1$ coins follow the given rules, as if they were all the coins.
- If a configuration starts with $T$ and ends with $H$, the middle $n-2$ coins follow the given rules, as if they were all the coins, until they are all $T$. After that, there are $2 n-1$ more steps: first coins $1,2, \ldots, n-1$ are turned over in that order, then coins $n, n-1, \ldots, 1$ are turned over in that order.

As this covers all configurations, and the number of steps is clearly finite for 0 or 1 coins, it follows by induction on $n$ that the number of steps is always finite.

We define $E_{A B}(n)$, where $A$ and $B$ are each one of $H, T$ or *, to be the average number of steps over configurations of length $n$ restricted to those that start with $A$, if $A$ is not *, and that end with $B$, if $B$ is not * (so * represents "either $H$ or $T$ "). The above observations tell us that, for $n \geqslant 2$ :

- $E_{H *}(n)=E(n-1)+1$.
- $E_{* T}(n)=E(n-1)$.
- $E_{H T}(n)=E(n-2)+1$ (by using both the observations for $H *$ and for $* T$ ).
- $E_{T H}(n)=E(n-2)+2 n-1$.

Now $E_{H *}(n)=\frac{1}{2}\left(E_{H H}(n)+E_{H T}(n)\right)$, so $E_{H H}(n)=2 E(n-1)-E(n-2)+1$. Similarly, $E_{T T}(n)=2 E(n-1)-E(n-2)-1$. So

$$
E(n)=\frac{1}{4}\left(E_{H T}(n)+E_{H H}(n)+E_{T T}(n)+E_{T H}(n)\right)=E(n-1)+\frac{n}{2} .
$$

We have $E(0)=0$ and $E(1)=\frac{1}{2}$, so by induction on $n$ we have $E(n)=\frac{1}{4} n(n+1)$.
Solution 3. Let $H_{i}$ be the number of heads in positions 1 to $i$ inclusive (so $H_{n}$ is the total number of heads), and let $I_{i}$ be 1 if the $i^{\text {th }}$ coin is a head, 0 otherwise. Consider the function

$$
t(i)=I_{i}+2\left(\min \left\{i, H_{n}\right\}-H_{i}\right) .
$$

We claim that $t(i)$ is the total number of times coin $i$ is turned over (which implies that the process terminates). Certainly $t(i)=0$ when all coins are tails, and $t(i)$ is always a nonnegative integer, so it suffices to show that when the $k^{\text {th }}$ coin is turned over (where $k=H_{n}$ ), $t(k)$ goes down by 1 and all the other $t(i)$ are unchanged. We show this by splitting into cases:

- If $i<k, I_{i}$ and $H_{i}$ are unchanged, and $\min \left\{i, H_{n}\right\}=i$ both before and after the coin flip, so $t(i)$ is unchanged.
- If $i>k, \min \left\{i, H_{n}\right\}=H_{n}$ both before and after the coin flip, and both $H_{n}$ and $H_{i}$ change by the same amount, so $t(i)$ is unchanged.
- If $i=k$ and the coin is heads, $I_{i}$ goes down by 1 , as do both $\min \left\{i, H_{n}\right\}=H_{n}$ and $H_{i}$; so $t(i)$ goes down by 1 .
- If $i=k$ and the coin is tails, $I_{i}$ goes up by $1, \min \left\{i, H_{n}\right\}=i$ is unchanged and $H_{i}$ goes up by 1 ; so $t(i)$ goes down by 1 .

We now need to compute the average value of

$$
\sum_{i=1}^{n} t(i)=\sum_{i=1}^{n} I_{i}+2 \sum_{i=1}^{n} \min \left\{i, H_{n}\right\}-2 \sum_{i=1}^{n} H_{i} .
$$

The average value of the first term is $\frac{1}{2} n$, and that of the third term is $-\frac{1}{2} n(n+1)$. To compute the second term, we sum over choices for the total number of heads, and then over the possible values of $i$, getting

$$
2^{1-n} \sum_{j=0}^{n}\binom{n}{j} \sum_{i=1}^{n} \min \{i, j\}=2^{1-n} \sum_{j=0}^{n}\binom{n}{j}\left(n j-\binom{j}{2}\right) .
$$

Now, in terms of trinomial coefficients,

$$
\sum_{j=0}^{n} j\binom{n}{j}=\sum_{j=1}^{n}\binom{n}{n-j, j-1,1}=n \sum_{j=0}^{n-1}\binom{n-1}{j}=2^{n-1} n
$$

and

$$
\sum_{j=0}^{n}\binom{j}{2}\binom{n}{j}=\sum_{j=2}^{n}\binom{n}{n-j, j-2,2}=\binom{n}{2} \sum_{j=0}^{n-2}\binom{n-2}{j}=2^{n-2}\binom{n}{2} .
$$

So the second term above is

$$
2^{1-n}\left(2^{n-1} n^{2}-2^{n-2}\binom{n}{2}\right)=n^{2}-\frac{n(n-1)}{4}
$$

and the required average is

$$
E(n)=\frac{1}{2} n+n^{2}-\frac{n(n-1)}{4}-\frac{1}{2} n(n+1)=\frac{n(n+1)}{4} .
$$

Solution 4. Harry has built a Turing machine to flip the coins for him. The machine is initially positioned at the $k^{\text {th }}$ coin, where there are $k$ heads (and the position before the first coin is considered to be the $0^{\text {th }}$ coin). The machine then moves according to the following rules, stopping when it reaches the position before the first coin: if the coin at its current position is $H$, it flips the coin and moves to the previous coin, while if the coin at its current position is $T$, it flips the coin and moves to the next position.

Consider the maximal sequences of consecutive moves in the same direction. Suppose the machine has $a$ consecutive moves to the next coin, before a move to the previous coin. After those $a$ moves, the $a$ coins flipped in those moves are all heads, as is the coin the machine is now at, so at least the next $a+1$ moves will all be moves to the previous coin. Similarly, $a$ consecutive moves to the previous coin are followed by at least $a+1$ consecutive moves to
the next coin. There cannot be more than $n$ consecutive moves in the same direction, so this proves that the process terminates (with a move from the first coin to the position before the first coin).

Thus we have a (possibly empty) sequence $a_{1}<\cdots<a_{t} \leqslant n$ giving the lengths of maximal sequences of consecutive moves in the same direction, where the final $a_{t}$ moves must be moves to the previous coin, ending before the first coin. We claim there is a bijection between initial configurations of the coins and such sequences. This gives

$$
E(n)=\frac{1}{2}(1+2+\cdots+n)=\frac{n(n+1)}{4}
$$

as required, since each $i$ with $1 \leqslant i \leqslant n$ will appear in half of the sequences, and will contribute $i$ to the number of moves when it does.

To see the bijection, consider following the sequence of moves backwards, starting with the machine just before the first coin and all coins showing tails. This certainly determines a unique configuration of coins that could possibly correspond to the given sequence. Furthermore, every coin flipped as part of the $a_{j}$ consecutive moves is also flipped as part of all subsequent sequences of $a_{k}$ consecutive moves, for all $k>j$, meaning that, as we follow the moves backwards, each coin is always in the correct state when flipped to result in a move in the required direction. (Alternatively, since there are $2^{n}$ possible configurations of coins and $2^{n}$ possible such ascending sequences, the fact that the sequence of moves determines at most one configuration of coins, and thus that there is an injection from configurations of coins to such ascending sequences, is sufficient for it to be a bijection, without needing to show that coins are in the right state as we move backwards.)

Solution 5. We explicitly describe what happens with an arbitrary sequence $C$ of $n$ coins. Suppose that $C$ contain $k$ heads at positions $1 \leqslant c_{1}<c_{2}<\cdots<c_{k} \leqslant n$.

Let $i$ be the minimal index such that $c_{i} \geqslant k$. Then the first few steps will consist of turning over the $k^{\mathrm{th}},(k+1)^{\mathrm{th}}, \ldots, c_{i}^{\mathrm{th}},\left(c_{i}-1\right)^{\mathrm{th}},\left(c_{i}-2\right)^{\mathrm{th}}, \ldots, k^{\mathrm{th}}$ coins in this order. After that we get a configuration with $k-1$ heads at the same positions as in the initial one, except for $c_{i}$. This part of the process takes $2\left(c_{i}-k\right)+1$ steps.

After that, the process acts similarly; by induction on the number of heads we deduce that the process ends. Moreover, if the $c_{i}$ disappear in order $c_{i_{1}}, \ldots, c_{i_{k}}$, the whole process takes

$$
\ell(C)=\sum_{j=1}^{k}\left(2\left(c_{i_{j}}-(k+1-j)\right)+1\right)=2 \sum_{j=1}^{k} c_{j}-2 \sum_{j=1}^{k}(k+1-j)+k=2 \sum_{j=1}^{k} c_{j}-k^{2}
$$

steps.
Now let us find the total value $S_{k}$ of $\ell(C)$ over all $\binom{n}{k}$ configurations with exactly $k$ heads. To sum up the above expression over those, notice that each number $1 \leqslant i \leqslant n$ appears as $c_{j}$ exactly $\binom{n-1}{k-1}$ times. Thus

$$
\begin{aligned}
S_{k}=2\binom{n-1}{k-1} & \sum_{i=1}^{n} i-\binom{n}{k} k^{2}=2 \frac{(n-1) \cdots(n-k+1)}{(k-1)!} \cdot \frac{n(n+1)}{2}-\frac{n \cdots(n-k+1)}{k!} k^{2} \\
& =\frac{n(n-1) \cdots(n-k+1)}{(k-1)!}((n+1)-k)=n(n-1)\binom{n-2}{k-1}+n\binom{n-1}{k-1} .
\end{aligned}
$$

Therefore, the total value of $\ell(C)$ over all configurations is

$$
\sum_{k=1}^{n} S_{k}=n(n-1) \sum_{k=1}^{n}\binom{n-2}{k-1}+n \sum_{k=1}^{n}\binom{n-1}{k-1}=n(n-1) 2^{n-2}+n 2^{n-1}=2^{n} \frac{n(n+1)}{4}
$$

Hence the required average is $E(n)=\frac{n(n+1)}{4}$.

C4. On a flat plane in Camelot, King Arthur builds a labyrinth $\mathfrak{L}$ consisting of $n$ walls, each of which is an infinite straight line. No two walls are parallel, and no three walls have a common point. Merlin then paints one side of each wall entirely red and the other side entirely blue.

At the intersection of two walls there are four corners: two diagonally opposite corners where a red side and a blue side meet, one corner where two red sides meet, and one corner where two blue sides meet. At each such intersection, there is a two-way door connecting the two diagonally opposite corners at which sides of different colours meet.

After Merlin paints the walls, Morgana then places some knights in the labyrinth. The knights can walk through doors, but cannot walk through walls.

Let $k(\mathfrak{L})$ be the largest number $k$ such that, no matter how Merlin paints the labyrinth $\mathfrak{L}$, Morgana can always place at least $k$ knights such that no two of them can ever meet. For each $n$, what are all possible values for $k(\mathfrak{L})$, where $\mathfrak{L}$ is a labyrinth with $n$ walls?
(Canada)

Answer: The only possible value of $k$ is $k=n+1$, no matter what shape the labyrinth is.

Solution 1. First we show by induction that the $n$ walls divide the plane into $\binom{n+1}{2}+1$ regions. The claim is true for $n=0$ as, when there are no walls, the plane forms a single region. When placing the $n^{\text {th }}$ wall, it intersects each of the $n-1$ other walls exactly once and hence splits each of $n$ of the regions formed by those other walls into two regions. By the induction hypothesis, this yields $\left(\binom{n}{2}+1\right)+n=\binom{n+1}{2}+1$ regions, proving the claim.

Now let $G$ be the graph with vertices given by the $\binom{n+1}{2}+1$ regions, and with two regions connected by an edge if there is a door between them.

We now show that no matter how Merlin paints the $n$ walls, Morgana can place at least $n+1$ knights. No matter how the walls are painted, there are exactly $\binom{n}{2}$ intersection points, each of which corresponds to a single edge in $G$. Consider adding the edges of $G$ sequentially and note that each edge reduces the number of connected components by at most one. Therefore the number of connected components of G is at least $\binom{n+1}{2}+1-\binom{n}{2}=n+1$. If Morgana places a knight in regions corresponding to different connected components of $G$, then no two knights can ever meet.

Now we give a construction showing that, no matter what shape the labyrinth is, Merlin can colour it such that there are exactly $n+1$ connected components, allowing Morgana to place at most $n+1$ knights.

First, we choose a coordinate system on the labyrinth so that none of the walls run due north-south, or due east-west. We then have Merlin paint the west face of each wall red, and the east face of each wall blue. We label the regions according to how many walls the region is on the east side of: the labels are integers between 0 and $n$.

We claim that, for each $i$, the regions labelled $i$ are connected by doors. First, we note that for each $i$ with $0 \leqslant i \leqslant n$ there is a unique region labelled $i$ which is unbounded to the north.

Now, consider a knight placed in some region with label $i$, and ask them to walk north (moving east or west by following the walls on the northern sides of regions, as needed). This knight will never get stuck: each region is convex, and so, if it is bounded to the north, it has a single northernmost vertex with a door northwards to another region with label $i$.

Eventually it will reach a region which is unbounded to the north, which will be the unique such region with label $i$. Hence every region with label $i$ is connected to this particular region, and so all regions with label $i$ are connected to each other.

As a result, there are exactly $n+1$ connected components, and Morgana can place at most $n+1$ knights.

Comment. Variations on this argument exist: some of them capture more information, and some of them capture less information, about the connected components according to this system of numbering.

For example, it can be shown that the unbounded regions are numbered $0,1, \ldots, n-1, n, n-1, \ldots, 1$ as one cycles around them, that the regions labelled 0 and $n$ are the only regions in their connected components, and that each other connected component forms a single chain running between the two unbounded ones. It is also possible to argue that the regions are acyclic without revealing much about their structure.

Solution 2. We give another description of a strategy for Merlin to paint the walls so that Morgana can place no more than $n+1$ knights.

Merlin starts by building a labyrinth of $n$ walls of his own design. He places walls in turn with increasing positive gradients, placing each so far to the right that all intersection points of previously-placed lines lie to the left of it. He paints each in such a way that blue is on the left and red is on the right.

For example, here is a possible sequence of four such lines $\ell_{1}, \ell_{2}, \ell_{3}, \ell_{4}$ :


We say that a region is "on the right" if it has $x$-coordinate unbounded above (note that if we only have one wall, then both regions are on the right). We claim inductively that, after placing $n$ lines, there are $n+1$ connected components in the resulting labyrinth, each of which contains exactly one region on the right. This is certainly true after placing 0 lines, as then there is only one region (and hence one connected component) and it is on the right.

When placing the $n^{\text {th }}$ line, it then cuts every one of the $n-1$ previously placed lines, and since it is to the right of all intersection points, the regions it cuts are exactly the $n$ regions on the right.


The addition of this line leaves all previous connected components with exactly one region on the right, and creates a new connected component containing exactly one region, and that region is also on the right. As a result, by induction, this particular labyrinth will have $n+1$ connected components.

Having built this labyrinth, Merlin then moves the walls one-by-one (by a sequence of continuous translations and rotations of lines) into the proper position of the given labyrinth, in such a way that no two lines ever become parallel.

The only time the configuration is changed is when one wall is moved through an intersection point of two others:


Note that all moves really do switch between two configurations like this: all sets of three lines have this colour configuration initially, and the rules on rotations mean they are preserved (in particular, we cannot create three lines creating a triangle with three red edges inwards, or three blue edges inwards).

However, as can be seen, such a move preserves the number of connected components, so in the painting this provides for Arthur's actual labyrinth, Morgana can still only place at most $n+1$ knights.

Comment. While these constructions are superficially distinct, they in fact result in the same colourings for any particular labyrinth. In fact, using the methods of Solution 2, it is possible to show that these are the only colourings that result in exactly $n+1$ connected components.

C5. On a certain social network, there are 2019 users, some pairs of which are friends, where friendship is a symmetric relation. Initially, there are 1010 people with 1009 friends each and 1009 people with 1010 friends each. However, the friendships are rather unstable, so events of the following kind may happen repeatedly, one at a time:

Let $A, B$, and $C$ be people such that $A$ is friends with both $B$ and $C$, but $B$ and $C$ are not friends; then $B$ and $C$ become friends, but $A$ is no longer friends with them.

Prove that, regardless of the initial friendships, there exists a sequence of such events after which each user is friends with at most one other user.

Common remarks. The problem has an obvious rephrasing in terms of graph theory. One is given a graph $G$ with 2019 vertices, 1010 of which have degree 1009 and 1009 of which have degree 1010. One is allowed to perform operations on $G$ of the following kind:

Suppose that vertex $A$ is adjacent to two distinct vertices $B$ and $C$ which are not adjacent to each other. Then one may remove the edges $A B$ and $A C$ from $G$ and add the edge $B C$ into $G$.

Call such an operation a refriending. One wants to prove that, via a sequence of such refriendings, one can reach a graph which is a disjoint union of single edges and vertices.

All of the solutions presented below will use this reformulation.
Solution 1. Note that the given graph is connected, since the total degree of any two vertices is at least 2018 and hence they are either adjacent or have at least one neighbour in common. Hence the given graph satisfies the following condition:

Every connected component of $G$ with at least three vertices is not complete and has a vertex of odd degree.

We will show that if a graph $G$ satisfies condition (1) and has a vertex of degree at least 2 , then there is a refriending on $G$ that preserves condition (1). Since refriendings decrease the total number of edges of $G$, by using a sequence of such refriendings, we must reach a graph $G$ with maximal degree at most 1 , so we are done.


Pick a vertex $A$ of degree at least 2 in a connected component $G^{\prime}$ of $G$. Since no component of $G$ with at least three vertices is complete we may assume that not all of the neighbours of $A$ are adjacent to one another. (For example, pick a maximal complete subgraph $K$ of $G^{\prime}$. Some vertex $A$ of $K$ has a neighbour outside $K$, and this neighbour is not adjacent to every vertex of $K$ by maximality.) Removing $A$ from $G$ splits $G^{\prime}$ into smaller connected components $G_{1}, \ldots, G_{k}$ (possibly with $k=1$ ), to each of which $A$ is connected by at least one edge. We divide into several cases.

Case 1: $k \geqslant 2$ and $A$ is connected to some $G_{i}$ by at least two edges.
Choose a vertex $B$ of $G_{i}$ adjacent to $A$, and a vertex $C$ in another component $G_{j}$ adjacent to $A$. The vertices $B$ and $C$ are not adjacent, and hence removing edges $A B$ and $A C$ and adding in edge $B C$ does not disconnect $G^{\prime}$. It is easy to see that this preserves the condition, since the refriending does not change the parity of the degrees of vertices.

Case 2: $k \geqslant 2$ and $A$ is connected to each $G_{i}$ by exactly one edge.
Consider the induced subgraph on any $G_{i}$ and the vertex $A$. The vertex $A$ has degree 1 in this subgraph; since the number of odd-degree vertices of a graph is always even, we see that $G_{i}$ has a vertex of odd degree (in $G$ ). Thus if we let $B$ and $C$ be any distinct neighbours of $A$, then removing edges $A B$ and $A C$ and adding in edge $B C$ preserves the above condition: the refriending creates two new components, and if either of these components has at least three vertices, then it cannot be complete and must contain a vertex of odd degree (since each $G_{i}$ does).

Case 3: $k=1$ and $A$ is connected to $G_{1}$ by at least three edges.
By assumption, $A$ has two neighbours $B$ and $C$ which are not adjacent to one another. Removing edges $A B$ and $A C$ and adding in edge $B C$ does not disconnect $G^{\prime}$. We are then done as in Case 1.

Case 4: $k=1$ and $A$ is connected to $G_{1}$ by exactly two edges.
Let $B$ and $C$ be the two neighbours of $A$, which are not adjacent. Removing edges $A B$ and $A C$ and adding in edge $B C$ results in two new components: one consisting of a single vertex; and the other containing a vertex of odd degree. We are done unless this second component would be a complete graph on at least 3 vertices. But in this case, $G_{1}$ would be a complete graph minus the single edge $B C$, and hence has at least 4 vertices since $G^{\prime}$ is not a 4 -cycle. If we let $D$ be a third vertex of $G_{1}$, then removing edges $B A$ and $B D$ and adding in edge $A D$ does not disconnect $G^{\prime}$. We are then done as in Case 1 .


Comment. In fact, condition 1 above precisely characterises those graphs which can be reduced to a graph of maximal degree $\leqslant 1$ by a sequence of refriendings.

Solution 2. As in the previous solution, note that a refriending preserves the property that a graph has a vertex of odd degree and (trivially) the property that it is not complete; note also that our initial graph is connected. We describe an algorithm to reduce our initial graph to a graph of maximal degree at most 1 , proceeding in two steps.

Step 1: There exists a sequence of refriendings reducing the graph to a tree.
Proof. Since the number of edges decreases with each refriending, it suffices to prove the following: as long as the graph contains a cycle, there exists a refriending such that the resulting graph is still connected. We will show that the graph in fact contains a cycle $Z$ and vertices $A, B, C$ such that $A$ and $B$ are adjacent in the cycle $Z, C$ is not in $Z$, and is adjacent to $A$ but not $B$. Removing edges $A B$ and $A C$ and adding in edge $B C$ keeps the graph connected, so we are done.


To find this cycle $Z$ and vertices $A, B, C$, we pursue one of two strategies. If the graph contains a triangle, we consider a largest complete subgraph $K$, which thus contains at least three vertices. Since the graph itself is not complete, there is a vertex $C$ not in $K$ connected to a vertex $A$ of $K$. By maximality of $K$, there is a vertex $B$ of $K$ not connected to $C$, and hence we are done by choosing a cycle $Z$ in $K$ through the edge $A B$.


If the graph is triangle-free, we consider instead a smallest cycle $Z$. This cycle cannot be Hamiltonian (i.e. it cannot pass through every vertex of the graph), since otherwise by minimality the graph would then have no other edges, and hence would have even degree at every vertex. We may thus choose a vertex $C$ not in $Z$ adjacent to a vertex $A$ of $Z$. Since the graph is triangle-free, it is not adjacent to any neighbour $B$ of $A$ in $Z$, and we are done.

Step 2: Any tree may be reduced to a disjoint union of single edges and vertices by a sequence of refriendings.

Proof. The refriending preserves the property of being acyclic. Hence, after applying a sequence of refriendings, we arrive at an acyclic graph in which it is impossible to perform any further refriendings. The maximal degree of any such graph is 1 : if it had a vertex $A$ with two neighbours $B, C$, then $B$ and $C$ would necessarily be nonadjacent since the graph is cycle-free, and so a refriending would be possible. Thus we reach a graph with maximal degree at most 1 as desired.

C6. Let $n>1$ be an integer. Suppose we are given $2 n$ points in a plane such that no three of them are collinear. The points are to be labelled $A_{1}, A_{2}, \ldots, A_{2 n}$ in some order. We then consider the $2 n$ angles $\angle A_{1} A_{2} A_{3}, \angle A_{2} A_{3} A_{4}, \ldots, \angle A_{2 n-2} A_{2 n-1} A_{2 n}, \angle A_{2 n-1} A_{2 n} A_{1}$, $\angle A_{2 n} A_{1} A_{2}$. We measure each angle in the way that gives the smallest positive value (i.e. between $0^{\circ}$ and $180^{\circ}$ ). Prove that there exists an ordering of the given points such that the resulting $2 n$ angles can be separated into two groups with the sum of one group of angles equal to the sum of the other group.

Comment. The first three solutions all use the same construction involving a line separating the points into groups of $n$ points each, but give different proofs that this construction works. Although Solution 1 is very short, the Problem Selection Committee does not believe any of the solutions is easy to find and thus rates this as a problem of medium difficulty.

Solution 1. Let $\ell$ be a line separating the points into two groups $(L$ and $R$ ) with $n$ points in each. Label the points $A_{1}, A_{2}, \ldots, A_{2 n}$ so that $L=\left\{A_{1}, A_{3}, \ldots, A_{2 n-1}\right\}$. We claim that this labelling works.

Take a line $s=A_{2 n} A_{1}$.
(a) Rotate $s$ around $A_{1}$ until it passes through $A_{2}$; the rotation is performed in a direction such that $s$ is never parallel to $\ell$.
(b) Then rotate the new $s$ around $A_{2}$ until it passes through $A_{3}$ in a similar manner.
(c) Perform $2 n-2$ more such steps, after which $s$ returns to its initial position.

The total (directed) rotation angle $\Theta$ of $s$ is clearly a multiple of $180^{\circ}$. On the other hand, $s$ was never parallel to $\ell$, which is possible only if $\Theta=0$. Now it remains to partition all the $2 n$ angles into those where $s$ is rotated anticlockwise, and the others.

Solution 2. When tracing a cyclic path through the $A_{i}$ in order, with straight line segments between consecutive points, let $\theta_{i}$ be the exterior angle at $A_{i}$, with a sign convention that it is positive if the path turns left and negative if the path turns right. Then $\sum_{i=1}^{2 n} \theta_{i}=360 k^{\circ}$ for some integer $k$. Let $\phi_{i}=\angle A_{i-1} A_{i} A_{i+1}($ indices $\bmod 2 n)$, defined as in the problem; thus $\phi_{i}=180^{\circ}-\left|\theta_{i}\right|$.

Let $L$ be the set of $i$ for which the path turns left at $A_{i}$ and let $R$ be the set for which it turns right. Then $S=\sum_{i \in L} \phi_{i}-\sum_{i \in R} \phi_{i}=(180(|L|-|R|)-360 k)^{\circ}$, which is a multiple of $360^{\circ}$ since the number of points is even. We will show that the points can be labelled such that $S=0$, in which case $L$ and $R$ satisfy the required condition of the problem.

Note that the value of $S$ is defined for a slightly larger class of configurations: it is OK for two points to coincide, as long as they are not consecutive, and OK for three points to be collinear, as long as $A_{i}, A_{i+1}$ and $A_{i+2}$ do not appear on a line in that order. In what follows it will be convenient, although not strictly necessary, to consider such configurations.

Consider how $S$ changes if a single one of the $A_{i}$ is moved along some straight-line path (not passing through any $A_{j}$ and not lying on any line $A_{j} A_{k}$, but possibly crossing such lines). Because $S$ is a multiple of $360^{\circ}$, and the angles change continuously, $S$ can only change when a point moves between $R$ and $L$. Furthermore, if $\phi_{j}=0$ when $A_{j}$ moves between $R$ and $L, S$ is unchanged; it only changes if $\phi_{j}=180^{\circ}$ when $A_{j}$ moves between those sets.

For any starting choice of points, we will now construct a new configuration, with labels such that $S=0$, that can be perturbed into the original one without any $\phi_{i}$ passing through $180^{\circ}$, so that $S=0$ for the original configuration with those labels as well.

Take some line such that there are $n$ points on each side of that line. The new configuration has $n$ copies of a single point on each side of the line, and a path that alternates between
sides of the line; all angles are 0 , so this configuration has $S=0$. Perturbing the points into their original positions, while keeping each point on its side of the line, no angle $\phi_{i}$ can pass through $180^{\circ}$, because no straight line can go from one side of the line to the other and back. So the perturbation process leaves $S=0$.

Comment. More complicated variants of this solution are also possible; for example, a path defined using four quadrants of the plane rather than just two half-planes.

Solution 3. First, let $\ell$ be a line in the plane such that there are $n$ points on one side and the other $n$ points on the other side. For convenience, assume $\ell$ is horizontal (otherwise, we can rotate the plane). Then we can use the terms "above", "below", "left" and "right" in the usual way. We denote the $n$ points above the line in an arbitrary order as $P_{1}, P_{2}, \ldots, P_{n}$, and the $n$ points below the line as $Q_{1}, Q_{2}, \ldots, Q_{n}$.

If we connect $P_{i}$ and $Q_{j}$ with a line segment, the line segment will intersect with the line $\ell$. Denote the intersection as $I_{i j}$. If $P_{i}$ is connected to $Q_{j}$ and $Q_{k}$, where $j<k$, then $I_{i j}$ and $I_{i k}$ are two different points, because $P_{i}, Q_{j}$ and $Q_{k}$ are not collinear.

Now we define a "sign" for each angle $\angle Q_{j} P_{i} Q_{k}$. Assume $j<k$. We specify that the sign is positive for the following two cases:

- if $i$ is odd and $I_{i j}$ is to the left of $I_{i k}$,
- if $i$ is even and $I_{i j}$ is to the right of $I_{i k}$.

Otherwise the sign of the angle is negative. If $j>k$, then the sign of $\angle Q_{j} P_{i} Q_{k}$ is taken to be the same as for $\angle Q_{k} P_{i} Q_{j}$.

Similarly, we can define the sign of $\angle P_{j} Q_{i} P_{k}$ with $j<k$ (or equivalently $\angle P_{k} Q_{i} P_{j}$ ). For example, it is positive when $i$ is odd and $I_{j i}$ is to the left of $I_{k i}$.

Henceforth, whenever we use the notation $\angle Q_{j} P_{i} Q_{k}$ or $\angle P_{j} Q_{i} P_{k}$ for a numerical quantity, it is understood to denote either the (geometric) measure of the angle or the negative of this measure, depending on the sign as specified above.

We now have the following important fact for signed angle measures:

$$
\begin{equation*}
\angle Q_{i_{1}} P_{k} Q_{i_{3}}=\angle Q_{i_{1}} P_{k} Q_{i_{2}}+\angle Q_{i_{2}} P_{k} Q_{i_{3}} \tag{1}
\end{equation*}
$$

for all points $P_{k}, Q_{i_{1}}, Q_{i_{2}}$ and $Q_{i_{3}}$ with $i_{1}<i_{2}<i_{3}$. The following figure shows a "natural" arrangement of the points. Equation (1) still holds for any other arrangement, as can be easily verified.


Similarly, we have

$$
\begin{equation*}
\angle P_{i_{1}} Q_{k} P_{i_{3}}=\angle P_{i_{1}} Q_{k} P_{i_{2}}+\angle P_{i_{2}} Q_{k} P_{i_{3}}, \tag{2}
\end{equation*}
$$

for all points $Q_{k}, P_{i_{1}}, P_{i_{2}}$ and $P_{i_{3}}$, with $i_{1}<i_{2}<i_{3}$.

We are now ready to specify the desired ordering $A_{1}, \ldots, A_{2 n}$ of the points:

- if $i \leqslant n$ is odd, put $A_{i}=P_{i}$ and $A_{2 n+1-i}=Q_{i}$;
- if $i \leqslant n$ is even, put $A_{i}=Q_{i}$ and $A_{2 n+1-i}=P_{i}$.

For example, for $n=3$ this ordering is $P_{1}, Q_{2}, P_{3}, Q_{3}, P_{2}, Q_{1}$. This sequence alternates between $P$ 's and $Q$ 's, so the above conventions specify a sign for each of the angles $A_{i-1} A_{i} A_{i+1}$. We claim that the sum of these $2 n$ signed angles equals 0 . If we can show this, it would complete the proof.

We prove the claim by induction. For brevity, we use the notation $\angle P_{i}$ to denote whichever of the $2 n$ angles has its vertex at $P_{i}$, and $\angle Q_{i}$ similarly.

First let $n=2$. If the four points can be arranged to form a convex quadrilateral, then the four line segments $P_{1} Q_{1}, P_{1} Q_{2}, P_{2} Q_{1}$ and $P_{2} Q_{2}$ constitute a self-intersecting quadrilateral. We use several figures to illustrate the possible cases.

The following figure is one possible arrangement of the points.


Then $\angle P_{1}$ and $\angle Q_{1}$ are positive, $\angle P_{2}$ and $\angle Q_{2}$ are negative, and we have

$$
\left|\angle P_{1}\right|+\left|\angle Q_{1}\right|=\left|\angle P_{2}\right|+\left|\angle Q_{2}\right| .
$$

With signed measures, we have

$$
\begin{equation*}
\angle P_{1}+\angle Q_{1}+\angle P_{2}+\angle Q_{2}=0 \tag{3}
\end{equation*}
$$

If we switch the labels of $P_{1}$ and $P_{2}$, we have the following picture:


Switching labels $P_{1}$ and $P_{2}$ has the effect of flipping the sign of all four angles (as well as swapping the magnitudes on the relabelled points); that is, the new values of ( $\angle P_{1}, \angle P_{2}, \angle Q_{1}, \angle Q_{2}$ ) equal the old values of ( $-\angle P_{2},-\angle P_{1},-\angle Q_{1},-\angle Q_{2}$ ). Consequently, equation (3) still holds. Similarly, when switching the labels of $Q_{1}$ and $Q_{2}$, or both the $P$ 's and the $Q$ 's, equation (3) still holds.

The remaining subcase of $n=2$ is that one point lies inside the triangle formed by the other three. We have the following picture.


We have

$$
\left|\angle P_{1}\right|+\left|\angle Q_{1}\right|+\left|\angle Q_{2}\right|=\left|\angle P_{2}\right| .
$$

and equation (3) holds.
Again, switching the labels for $P$ 's or the $Q$ 's will not affect the validity of equation (3). Also, if the point lying inside the triangle of the other three is one of the $Q$ 's rather than the $P$ 's, the result still holds, since our sign convention is preserved when we relabel $Q$ 's as $P$ 's and vice-versa and reflect across $\ell$.

We have completed the proof of the claim for $n=2$.
Assume the claim holds for $n=k$, and we wish to prove it for $n=k+1$. Suppose we are given our $2(k+1)$ points. First ignore $P_{k+1}$ and $Q_{k+1}$, and form $2 k$ angles from $P_{1}, \ldots, P_{k}$, $Q_{1}, \ldots, Q_{k}$ as in the $n=k$ case. By the induction hypothesis we have

$$
\sum_{i=1}^{k}\left(\angle P_{i}+\angle Q_{i}\right)=0
$$

When we add in the two points $P_{k+1}$ and $Q_{k+1}$, this changes our angles as follows:

- the angle at $P_{k}$ changes from $\angle Q_{k-1} P_{k} Q_{k}$ to $\angle Q_{k-1} P_{k} Q_{k+1}$;
- the angle at $Q_{k}$ changes from $\angle P_{k-1} Q_{k} P_{k}$ to $\angle P_{k-1} Q_{k} P_{k+1}$;
- two new angles $\angle Q_{k} P_{k+1} Q_{k+1}$ and $\angle P_{k} Q_{k+1} P_{k+1}$ are added.

We need to prove the changes have no impact on the total sum. In other words, we need to prove

$$
\begin{equation*}
\left(\angle Q_{k-1} P_{k} Q_{k+1}-\angle Q_{k-1} P_{k} Q_{k}\right)+\left(\angle P_{k-1} Q_{k} P_{k+1}-\angle P_{k-1} Q_{k} P_{k}\right)+\left(\angle P_{k+1}+\angle Q_{k+1}\right)=0 . \tag{4}
\end{equation*}
$$

In fact, from equations (1) and (2), we have

$$
\angle Q_{k-1} P_{k} Q_{k+1}-\angle Q_{k-1} P_{k} Q_{k}=\angle Q_{k} P_{k} Q_{k+1}
$$

and

$$
\angle P_{k-1} Q_{k} P_{k+1}-\angle P_{k-1} Q_{k} P_{k}=\angle P_{k} Q_{k} P_{k+1}
$$

Therefore, the left hand side of equation (4) becomes $\angle Q_{k} P_{k} Q_{k+1}+\angle P_{k} Q_{k} P_{k+1}+\angle Q_{k} P_{k+1} Q_{k+1}+$ $\angle P_{k} Q_{k+1} P_{k+1}$, which equals 0 , simply by applying the $n=2$ case of the claim. This completes the induction.

Solution 4. We shall think instead of the problem as asking us to assign a weight $\pm 1$ to each angle, such that the weighted sum of all the angles is zero.

Given an ordering $A_{1}, \ldots, A_{2 n}$ of the points, we shall assign weights according to the following recipe: walk in order from point to point, and assign the left turns +1 and the right turns -1 . This is the same weighting as in Solution 3, and as in that solution, the weighted sum is a multiple of $360^{\circ}$.

We now aim to show the following:
Lemma. Transposing any two consecutive points in the ordering changes the weighted sum by $\pm 360^{\circ}$ or 0 .

Knowing that, we can conclude quickly: if the ordering $A_{1}, \ldots, A_{2 n}$ has weighted angle sum $360 k^{\circ}$, then the ordering $A_{2 n}, \ldots, A_{1}$ has weighted angle sum $-360 k^{\circ}$ (since the angles are the same, but left turns and right turns are exchanged). We can reverse the ordering of $A_{1}$, $\ldots, A_{2 n}$ by a sequence of transpositions of consecutive points, and in doing so the weighted angle sum must become zero somewhere along the way.

We now prove that lemma:
Proof. Transposing two points amounts to taking a section $A_{k} A_{k+1} A_{k+2} A_{k+3}$ as depicted, reversing the central line segment $A_{k+1} A_{k+2}$, and replacing its two neighbours with the dotted lines.


Figure 1: Transposing two consecutive vertices: before (left) and afterwards (right)
In each triangle, we alter the sum by $\pm 180^{\circ}$. Indeed, using (anticlockwise) directed angles modulo $360^{\circ}$, we either add or subtract all three angles of each triangle.

Hence both triangles together alter the sum by $\pm 180 \pm 180^{\circ}$, which is $\pm 360^{\circ}$ or 0 .

C7. There are 60 empty boxes $B_{1}, \ldots, B_{60}$ in a row on a table and an unlimited supply of pebbles. Given a positive integer $n$, Alice and Bob play the following game.

In the first round, Alice takes $n$ pebbles and distributes them into the 60 boxes as she wishes. Each subsequent round consists of two steps:
(a) Bob chooses an integer $k$ with $1 \leqslant k \leqslant 59$ and splits the boxes into the two groups $B_{1}, \ldots, B_{k}$ and $B_{k+1}, \ldots, B_{60}$.
(b) Alice picks one of these two groups, adds one pebble to each box in that group, and removes one pebble from each box in the other group.

Bob wins if, at the end of any round, some box contains no pebbles. Find the smallest $n$ such that Alice can prevent Bob from winning.
(Czech Republic)
Answer: $n=960$. In general, if there are $N>1$ boxes, the answer is $n=\left\lfloor\frac{N}{2}+1\right\rfloor\left\lceil\frac{N}{2}+1\right\rceil-1$.
Common remarks. We present solutions for the general case of $N>1$ boxes, and write $M=\left\lfloor\frac{N}{2}+1\right\rfloor\left\lceil\frac{N}{2}+1\right\rceil-1$ for the claimed answer. For $1 \leqslant k<N$, say that Bob makes a $k$-move if he splits the boxes into a left group $\left\{B_{1}, \ldots, B_{k}\right\}$ and a right group $\left\{B_{k+1}, \ldots, B_{N}\right\}$. Say that one configuration dominates another if it has at least as many pebbles in each box, and say that it strictly dominates the other configuration if it also has more pebbles in at least one box. (Thus, if Bob wins in some configuration, he also wins in every configuration that it dominates.)

It is often convenient to consider ' V -shaped' configurations; for $1 \leqslant i \leqslant N$, let $V_{i}$ be the configuration where $B_{j}$ contains $1+|j-i|$ pebbles (i.e. where the $i^{\text {th }}$ box has a single pebble and the numbers increase by one in both directions, so the first box has $i$ pebbles and the last box has $N+1-i$ pebbles). Note that $V_{i}$ contains $\frac{1}{2} i(i+1)+\frac{1}{2}(N+1-i)(N+2-i)-1$ pebbles. If $i=\left\lceil\frac{N}{2}\right\rceil$, this number equals $M$.

Solutions split naturally into a strategy for Alice (starting with $M$ pebbles and showing she can prevent Bob from winning) and a strategy for Bob (showing he can win for any starting configuration with at most $M-1$ pebbles). The following observation is also useful to simplify the analysis of strategies for Bob.
Observation A. Consider two consecutive rounds. Suppose that in the first round Bob made a $k$-move and Alice picked the left group, and then in the second round Bob makes an $\ell$-move, with $\ell>k$. We may then assume, without loss of generality, that Alice again picks the left group.
Proof. Suppose Alice picks the right group in the second round. Then the combined effect of the two rounds is that each of the boxes $B_{k+1}, \ldots, B_{\ell}$ lost two pebbles (and the other boxes are unchanged). Hence this configuration is strictly dominated by that before the first round, and it suffices to consider only Alice's other response.

Solution 1 (Alice). Alice initially distributes pebbles according to $V_{\left\lceil\frac{N}{2}\right\rceil}$. Suppose the current configuration of pebbles dominates $V_{i}$. If Bob makes a $k$-move with $k \geqslant i$ then Alice picks the left group, which results in a configuration that dominates $V_{i+1}$. Likewise, if Bob makes a $k$-move with $k<i$ then Alice picks the right group, which results in a configuration that dominates $V_{i-1}$. Since none of $V_{1}, \ldots, V_{N}$ contains an empty box, Alice can prevent Bob from ever winning.

Solution 1 (Bob). The key idea in this solution is the following claim.
Claim. If there exist a positive integer $k$ such that there are at least $2 k$ boxes that have at most $k$ pebbles each then Bob can force a win.
Proof. We ignore the other boxes. First, Bob makes a $k$-move (splits the $2 k$ boxes into two groups of $k$ boxes each). Without loss of generality, Alice picks the left group. Then Bob makes a $(k+1)$-move, $\ldots$, a $(2 k-1)$-move. By Observation A, we may suppose Alice always picks the left group. After Bob's $(2 k-1)$-move, the rightmost box becomes empty and Bob wins.

Now, we claim that if $n<M$ then either there already exists an empty box, or there exist a positive integer $k$ and $2 k$ boxes with at most $k$ pebbles each (and thus Bob can force a win). Otherwise, assume each box contains at least 1 pebble, and for each $1 \leqslant k \leqslant\left\lfloor\frac{N}{2}\right\rfloor$, at least $N-(2 k-1)=N+1-2 k$ boxes contain at least $k+1$ pebbles. Summing, there are at least as many pebbles in total as in $V_{\left\lceil\frac{N}{2}\right\rceil}$; that is, at least $M$ pebbles, as desired.

Solution 2 (Alice). Let $K=\left\lfloor\frac{N}{2}+1\right\rfloor$. Alice starts with the boxes in the configuration $V_{K}$. For each of Bob's $N-1$ possible choices, consider the subset of rounds in which he makes that choice. In that subset of rounds, Alice alternates between picking the left group and picking the right group; the first time Bob makes that choice, Alice picks the group containing the $K^{\text {th }}$ box. Thus, at any time during the game, the number of pebbles in each box depends only on which choices Bob has made an odd number of times. This means that the number of pebbles in a box could decrease by at most the number of choices for which Alice would have started by removing a pebble from the group containing that box. These numbers are, for each box,

$$
\left\lfloor\frac{N}{2}\right\rfloor,\left\lfloor\frac{N}{2}-1\right\rfloor, \ldots, 1,0,1, \ldots,\left\lceil\frac{N}{2}-1\right\rceil .
$$

These are pointwise less than the numbers of pebbles the boxes started with, meaning that no box ever becomes empty with this strategy.

Solution 2 (Bob). Let $K=\left\lfloor\frac{N}{2}+1\right\rfloor$. For Bob's strategy, we consider a configuration $X$ with at most $M-1$ pebbles, and we make use of Observation A. Consider two configurations with $M$ pebbles: $V_{K}$ and $V_{N+1-K}$ (if $n$ is odd, they are the same configuration; if $n$ is even, one is the reverse of the other). The configuration $X$ has fewer pebbles than $V_{K}$ in at least one box, and fewer pebbles than $V_{N+1-K}$ in at least one box.

Suppose first that, with respect to one of those configurations (without loss of generality $V_{K}$ ), $X$ has fewer pebbles in one of the boxes in the half where they have $1,2, \ldots,\left\lceil\frac{N}{2}\right\rceil$ pebbles (the right half in $V_{K}$ if $N$ is even; if $N$ is odd, we can take it to be the right half, without loss of generality, as the configuration is symmetric). Note that the number cannot be fewer in the box with 1 pebble in $V_{K}$, because then it would have 0 pebbles. Bob then does a $K$-move. If Alice picks the right group, the total number of pebbles goes down and we restart Bob's strategy with a smaller number of pebbles. If Alice picks the left group, Bob follows with a $(K+1)$-move, a $(K+2)$-move, and so on; by Observation A we may assume Alice always picks the left group. But whichever box in the right half had fewer pebbles in $X$ than in $V_{K}$ ends up with 0 pebbles at some point in this sequence of moves.

Otherwise, $N$ is even, and for both of those configurations, there are fewer pebbles in $X$ only on the $2,3, \ldots, \frac{N}{2}+1$ side. That is, the numbers of pebbles in $X$ are at least

$$
\begin{equation*}
\frac{N}{2}, \frac{N}{2}-1, \ldots, 1,1, \ldots, \frac{N}{2} \tag{C}
\end{equation*}
$$

with equality occurring at least once on each side. Bob does an $\frac{N}{2}$-move. Whichever group Alice chooses, the total number of pebbles is unchanged, and the side from which pebbles are removed now has a box with fewer pebbles than in $(C)$, so the previous case of Bob's strategy can now be applied.

Solution 3 (Bob). For any configuration $C$, define $L(C)$ to be the greatest integer such that, for all $0 \leqslant i \leqslant N-1$, the box $B_{i+1}$ contains at least $L(C)-i$ pebbles. Similarly, define $R(C)$ to be greatest integer such that, for all $0 \leqslant i \leqslant N-1$, the box $B_{N-i}$ contains at least $R(C)-i$ pebbles. (Thus, $C$ dominates the 'left half' of $V_{L(C)}$ and the 'right half' of $V_{N+1-R(C)}$.) Then $C$ dominates a ' V -shaped' configuration if and only if $L(C)+R(C) \geqslant N+1$. Note that if $C$ dominates a $V$-shaped configuration, it has at least $M$ pebbles.

Now suppose that there are fewer than $M$ pebbles, so we have $L(C)+R(C) \leqslant N$. Then Bob makes an $L(C)$-move (or more generally any move with at least $L(C)$ boxes on the left and $R(C)$ boxes on the right). Let $C^{\prime}$ be the new configuration, and suppose that no box becomes empty (otherwise Bob has won). If Alice picks the left group, we have $L\left(C^{\prime}\right)=L(C)+1$ and $R\left(C^{\prime}\right)=R(C)-1$. Otherwise, we have $L\left(C^{\prime}\right)=L(C)-1$ and $R\left(C^{\prime}\right)=R(C)+1$. In either case, we have $L\left(C^{\prime}\right)+R\left(C^{\prime}\right) \leqslant N$.

Bob then repeats this strategy, until one of the boxes becomes empty. Since the condition in Observation A holds, we may assume that Alice picks a group on the same side each time. Then one of $L$ and $R$ is strictly decreasing; without loss of generality assume that $L$ strictly decreases. At some point we reach $L=1$. If $B_{2}$ is still nonempty, then $B_{1}$ must contain a single pebble. Bob makes a 1 -move, and by Observation A, Alice must (eventually) pick the right group, making this box empty.

C8. Alice has a map of Wonderland, a country consisting of $n \geqslant 2$ towns. For every pair of towns, there is a narrow road going from one town to the other. One day, all the roads are declared to be "one way" only. Alice has no information on the direction of the roads, but the King of Hearts has offered to help her. She is allowed to ask him a number of questions. For each question in turn, Alice chooses a pair of towns and the King of Hearts tells her the direction of the road connecting those two towns.

Alice wants to know whether there is at least one town in Wonderland with at most one outgoing road. Prove that she can always find out by asking at most $4 n$ questions.

Comment. This problem could be posed with an explicit statement about points being awarded for weaker bounds $c n$ for some $c>4$, in the style of IMO 2014 Problem 6.
(Thailand)
Solution. We will show Alice needs to ask at most $4 n-7$ questions. Her strategy has the following phases. In what follows, $S$ is the set of towns that Alice, so far, does not know to have more than one outgoing road (so initially $|S|=n$ ).

Phase 1. Alice chooses any two towns, say $A$ and $B$. Without loss of generality, suppose that the King of Hearts' answer is that the road goes from $A$ to $B$.

At the end of this phase, Alice has asked 1 question.
Phase 2. During this phase there is a single (variable) town $T$ that is known to have at least one incoming road but not yet known to have any outgoing roads. Initially, $T$ is $B$. Alice does the following $n-2$ times: she picks a town $X$ she has not asked about before, and asks the direction of the road between $T$ and $X$. If it is from $X$ to $T, T$ is unchanged; if it is from $T$ to $X, X$ becomes the new choice of town $T$, as the previous $T$ is now known to have an outgoing road.

At the end of this phase, Alice has asked a total of $n-1$ questions. The final town $T$ is not yet known to have any outgoing roads, while every other town has exactly one outgoing road known. The undirected graph of roads whose directions are known is a tree.

Phase 3. During this phase, Alice asks about the directions of all roads between $T$ and another town she has not previously asked about, stopping if she finds two outgoing roads from $T$. This phase involves at most $n-2$ questions. If she does not find two outgoing roads from $T$, she has answered her original question with at most $2 n-3 \leqslant 4 n-7$ questions, so in what follows we suppose that she does find two outgoing roads, asking a total of $k$ questions in this phase, where $2 \leqslant k \leqslant n-2$ (and thus $n \geqslant 4$ for what follows).

For every question where the road goes towards $T$, the town at the other end is removed from $S$ (as it already had one outgoing road known), while the last question resulted in $T$ being removed from $S$. So at the end of this phase, $|S|=n-k+1$, while a total of $n+k-1$ questions have been asked. Furthermore, the undirected graph of roads within $S$ whose directions are known contains no cycles (as $T$ is no longer a member of $S$, all questions asked in this phase involved $T$ and the graph was a tree before this phase started). Every town in $S$ has exactly one outgoing road known (not necessarily to another town in $S$ ).

Phase 4. During this phase, Alice repeatedly picks any pair of towns in $S$ for which she does not know the direction of the road between them. Because every town in $S$ has exactly one outgoing road known, this always results in the removal of one of those two towns from $S$. Because there are no cycles in the graph of roads of known direction within $S$, this can continue until there are at most 2 towns left in $S$.

If it ends with $t$ towns left, $n-k+1-t$ questions were asked in this phase, so a total of $2 n-t$ questions have been asked.

Phase 5. During this phase, Alice asks about all the roads from the remaining towns in $S$ that she has not previously asked about. She has definitely already asked about any road between those towns (if $t=2$ ). She must also have asked in one of the first two phases about
at least one other road involving one of those towns (as those phases resulted in a tree with $n>2$ vertices). So she asks at most $t(n-t)-1$ questions in this phase.

At the end of this phase, Alice knows whether any town has at most one outgoing road. If $t=1$, at most $3 n-3 \leqslant 4 n-7$ questions were needed in total, while if $t=2$, at most $4 n-7$ questions were needed in total.

Comment 1. The version of this problem originally submitted asked only for an upper bound of $5 n$, which is much simpler to prove. The Problem Selection Committee preferred a version with an asymptotically optimal constant. In the following comment, we will show that the constant is optimal.

Comment 2. We will show that Alice cannot always find out by asking at most $4 n-3\left(\log _{2} n\right)-$ 15 questions, if $n \geqslant 8$.

To show this, we suppose the King of Hearts is choosing the directions as he goes along, only picking the direction of a road when Alice asks about it for the first time. We provide a strategy for the King of Hearts that ensures that, after the given number of questions, the map is still consistent both with the existence of a town with at most one outgoing road, and with the nonexistence of such a town. His strategy has the following phases. When describing how the King of Hearts' answer to a question is determined below, we always assume he is being asked about a road for the first time (otherwise, he just repeats his previous answer for that road). This strategy is described throughout in graph-theoretic terms (vertices and edges rather than towns and roads).

Phase 1. In this phase, we consider the undirected graph formed by edges whose directions are known. The phase terminates when there are exactly 8 connected components whose undirected graphs are trees. The following invariant is maintained: in a component with $k$ vertices whose undirected graph is a tree, every vertex has at most $\left\lfloor\log _{2} k\right\rfloor$ edges into it.

- If the King of Hearts is asked about an edge between two vertices in the same component, or about an edge between two components at least one of which is not a tree, he chooses any direction for that edge arbitrarily.
- If he is asked about an edge between a vertex in component $A$ that has $a$ vertices and is a tree and a vertex in component $B$ that has $b$ vertices and is a tree, suppose without loss of generality that $a \geqslant b$. He then chooses the edge to go from $A$ to $B$. In this case, the new number of edges into any vertex is at most $\max \left\{\left\lfloor\log _{2} a\right\rfloor,\left\lfloor\log _{2} b\right\rfloor+1\right\} \leqslant\left\lfloor\log _{2}(a+b)\right\rfloor$.

In all cases, the invariant is preserved, and the number of tree components either remains unchanged or goes down by 1. Assuming Alice does not repeat questions, the process must eventually terminate with 8 tree components, and at least $n-8$ questions having been asked.

Note that each tree component contains at least one vertex with no outgoing edges. Colour one such vertex in each tree component red.

Phase 2. Let $V_{1}, V_{2}$ and $V_{3}$ be the three of the red vertices whose components are smallest (so their components together have at most $\left\lfloor\frac{3}{8} n\right\rfloor$ vertices, with each component having at most $\left\lfloor\frac{3}{8} n-2\right\rfloor$ vertices). Let sets $C_{1}, C_{2}, \ldots$ be the connected components after removing the $V_{j}$. By construction, there are no edges with known direction between $C_{i}$ and $C_{j}$ for $i \neq j$, and there are at least five such components.

If at any point during this phase, the King of Hearts is asked about an edge within one of the $C_{i}$, he chooses an arbitrary direction. If he is asked about an edge between $C_{i}$ and $C_{j}$ for $i \neq j$, he answers so that all edges go from $C_{i}$ to $C_{i+1}$ and $C_{i+2}$, with indices taken modulo the number of components, and chooses arbitrarily for other pairs. This ensures that all vertices other than the $V_{j}$ will have more than one outgoing edge.

For edges involving one of the $V_{j}$ he answers as follows, so as to remain consistent for as long as possible with both possibilities for whether one of those vertices has at most one outgoing edge. Note that as they were red vertices, they have no outgoing edges at the start of this phase. For edges between two of the $V_{j}$, he answers that the edges go from $V_{1}$ to $V_{2}$, from $V_{2}$ to $V_{3}$ and from $V_{3}$ to $V_{1}$. For edges between $V_{j}$ and some other vertex, he always answers that the edge goes into $V_{j}$, except for the last such edge for which he is asked the question for any given $V_{j}$, for which he answers that the
edge goes out of $V_{j}$. Thus, as long as at least one of the $V_{j}$ has not had the question answered for all the vertices that are not among the $V_{j}$, his answers are still compatible both with all vertices having more than one outgoing edge, and with that $V_{j}$ having only one outgoing edge.

At the start of this phase, each of the $V_{j}$ has at most $\left\lfloor\log _{2}\left\lfloor\frac{3}{8} n-2\right\rfloor\right\rfloor<\left(\log _{2} n\right)-1$ incoming edges. Thus, Alice cannot determine whether some vertex has only one outgoing edge within 3 ( $n-$ $\left.3-\left(\left(\log _{2} n\right)-1\right)\right)-1$ questions in this phase; that is, $4 n-3\left(\log _{2} n\right)-15$ questions total.

Comment 3. We can also improve the upper bound slightly, to $4 n-2\left(\log _{2} n\right)+1$. (We do not know where the precise minimum number of questions lies between $4 n-3\left(\log _{2} n\right)+O(1)$ and $4 n-2\left(\log _{2} n\right)+$ $O(1)$.) Suppose $n \geqslant 5$ (otherwise no questions are required at all).

To do this, we replace Phases 1 and 2 of the given solution with a different strategy that also results in a spanning tree where one vertex $V$ is not known to have any outgoing edges, and all other vertices have exactly one outgoing edge known, but where there is more control over the numbers of incoming edges. In Phases 3 and 4 we then take more care about the order in which pairs of towns are chosen, to ensure that each of the remaining towns has already had a question asked about at least $\log _{2} n+O(1)$ edges.

Define trees $T_{m}$ with $2^{m}$ vertices, exactly one of which (the root) has no outgoing edges and the rest of which have exactly one outgoing edge, as follows: $T_{0}$ is a single vertex, while $T_{m}$ is constructed by joining the roots of two copies of $T_{m-1}$ with an edge in either direction. If $n=2^{m}$ we can readily ask $n-1$ questions, resulting in a tree $T_{m}$ for the edges with known direction: first ask about $2^{m-1}$ disjoint pairs of vertices, then about $2^{m-2}$ disjoint pairs of the roots of the resulting $T_{1}$ trees, and so on. For the general case, where $n$ is not a power of 2 , after $k$ stages of this process we have $\left\lfloor n / 2^{k}\right\rfloor$ trees, each of which is like $T_{k}$ but may have some extra vertices (but, however, a unique root). If there are an even number of trees, then ask about pairs of their roots. If there are an odd number (greater than 1) of trees, when a single $T_{k}$ is left over, ask about its root together with that of one of the $T_{k+1}$ trees.

Say $m=\left\lfloor\log _{2} n\right\rfloor$. The result of that process is a single $T_{m}$ tree, possibly with some extra vertices but still a unique root $V$. That root has at least $m$ incoming edges, and we may list vertices $V_{0}$, $\ldots, V_{m-1}$ with edges to $V$, such that, for all $0 \leqslant i<m$, vertex $V_{i}$ itself has at least $i$ incoming edges.

Now divide the vertices other than $V$ into two parts: $A$ has all vertices at an odd distance from $V$ and $B$ has all the vertices at an even distance from $B$. Both $A$ and $B$ are nonempty; $A$ contains the $V_{i}$, while $B$ contains a sequence of vertices with at least $0,1, \ldots, m-2$ incoming edges respectively, similar to the $V_{i}$. There are no edges with known direction within $A$ or within $B$.

In Phase 3, then ask about edges between $V$ and other vertices: first those in $B$, in order of increasing number of incoming edges to the other vertex, then those in $A$, again in order of increasing number of incoming edges, which involves asking at most $n-1-m$ questions in this phase. If two outgoing edges are not found from $V$, at most $2 n-2-m \leqslant 4 n-2\left(\log _{2} n\right)+1$ questions needed to be asked in total, so we suppose that two outgoing edges were found, with $k$ questions asked in this phase, where $2 \leqslant k \leqslant n-1-m$. The state of $S$ is as described in the solution above, with the additional property that, since $S$ must still contain all vertices with edges to $V$, it contains the vertices $V_{i}$ described above.

In Phase 4, consider the vertices left in $B$, in increasing order of number of edges incoming to a vertex. If $s$ is the least number of incoming edges to such a vertex, then, for any $s \leqslant t \leqslant m-2$, there are at least $m-t-2$ vertices with more than $t$ incoming edges. Repeatedly asking about the pair of vertices left in $B$ with the least numbers of incoming edges results in a single vertex left over (if any were in $B$ at all at the start of this phase) with at least $m-2$ incoming edges. Doing the same with $A$ (which must be nonempty) leaves a vertex with at least $m-1$ incoming edges.

Thus if only $A$ is nonempty we ask at most $n-m$ questions in Phase 5 , so in total at most $3 n-m-1$ questions, while if both are nonempty we ask at most $2 n-2 m+1$ questions in Phase 5 , so in total at most $4 n-2 m-1<4 n-2\left(\log _{2} n\right)+1$ questions.

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C9. For any two different real numbers $x$ and $y$, we define $D(x, y)$ to be the unique integer $d$ satisfying $2^{d} \leqslant|x-y|<2^{d+1}$. Given a set of reals $\mathcal{F}$, and an element $x \in \mathcal{F}$, we say that the scales of $x$ in $\mathcal{F}$ are the values of $D(x, y)$ for $y \in \mathcal{F}$ with $x \neq y$.

Let $k$ be a given positive integer. Suppose that each member $x$ of $\mathcal{F}$ has at most $k$ different scales in $\mathcal{F}$ (note that these scales may depend on $x$ ). What is the maximum possible size of $\mathcal{F}$ ?
(Italy)
Answer: The maximum possible size of $\mathcal{F}$ is $2^{k}$.
Common remarks. For convenience, we extend the use of the word scale: we say that the scale between two reals $x$ and $y$ is $D(x, y)$.

Solution. We first construct a set $\mathcal{F}$ with $2^{k}$ members, each member having at most $k$ different scales in $\mathcal{F}$. Take $\mathcal{F}=\left\{0,1,2, \ldots, 2^{k}-1\right\}$. The scale between any two members of $\mathcal{F}$ is in the set $\{0,1, \ldots, k-1\}$.

We now show that $2^{k}$ is an upper bound on the size of $\mathcal{F}$. For every finite set $\mathcal{S}$ of real numbers, and every real $x$, let $r_{\mathcal{S}}(x)$ denote the number of different scales of $x$ in $\mathcal{S}$. That is, $r_{\mathcal{S}}(x)=|\{D(x, y): x \neq y \in \mathcal{S}\}|$. Thus, for every element $x$ of the set $\mathcal{F}$ in the problem statement, we have $r_{\mathcal{F}}(x) \leqslant k$. The condition $|\mathcal{F}| \leqslant 2^{k}$ is an immediate consequence of the following lemma.
Lemma. Let $\mathcal{S}$ be a finite set of real numbers, and define

$$
w(\mathcal{S})=\sum_{x \in \mathcal{S}} 2^{-r_{\mathcal{S}}(x)}
$$

Then $w(\mathcal{S}) \leqslant 1$.
Proof. Induction on $n=|\mathcal{S}|$. If $\mathcal{S}=\{x\}$, then $r_{\mathcal{S}}(x)=0$, so $w(\mathcal{S})=1$.
Assume now $n \geqslant 2$, and let $x_{1}<\cdots<x_{n}$ list the members of $\mathcal{S}$. Let $d$ be the minimal scale between two distinct elements of $\mathcal{S}$; then there exist neighbours $x_{t}$ and $x_{t+1}$ with $D\left(x_{t}, x_{t+1}\right)=d$. Notice that for any two indices $i$ and $j$ with $j-i>1$ we have $D\left(x_{i}, x_{j}\right)>d$, since

$$
\left|x_{i}-x_{j}\right|=\left|x_{i+1}-x_{i}\right|+\left|x_{j}-x_{i+1}\right| \geqslant 2^{d}+2^{d}=2^{d+1} .
$$

Now choose the minimal $i \leqslant t$ and the maximal $j \geqslant t+1$ such that $D\left(x_{i}, x_{i+1}\right)=$ $D\left(x_{i+1}, x_{i+2}\right)=\cdots=D\left(x_{j-1}, x_{j}\right)=d$.

Let $E$ be the set of all the $x_{s}$ with even indices $i \leqslant s \leqslant j, O$ be the set of those with odd indices $i \leqslant s \leqslant j$, and $R$ be the rest of the elements (so that $\mathcal{S}$ is the disjoint union of $E, O$ and $R$ ). Set $\mathcal{S}_{O}=R \cup O$ and $\mathcal{S}_{E}=R \cup E$; we have $\left|\mathcal{S}_{O}\right|<|\mathcal{S}|$ and $\left|\mathcal{S}_{E}\right|<|\mathcal{S}|$, so $w\left(\mathcal{S}_{O}\right), w\left(\mathcal{S}_{E}\right) \leqslant 1$ by the inductive hypothesis.

Clearly, $r_{\mathcal{S}_{O}}(x) \leqslant r_{\mathcal{S}}(x)$ and $r_{\mathcal{S}_{E}}(x) \leqslant r_{\mathcal{S}}(x)$ for any $x \in R$, and thus

$$
\begin{aligned}
\sum_{x \in R} 2^{-r \mathcal{S}(x)} & =\frac{1}{2} \sum_{x \in R}\left(2^{-r \mathcal{S}(x)}+2^{-r_{\mathcal{S}}(x)}\right) \\
& \leqslant \frac{1}{2} \sum_{x \in R}\left(2^{-r_{\mathcal{S}_{O}}(x)}+2^{-r_{\mathcal{S}_{E}}(x)}\right) .
\end{aligned}
$$

On the other hand, for every $x \in O$, there is no $y \in \mathcal{S}_{O}$ such that $D_{\mathcal{S}_{O}}(x, y)=d$ (as all candidates from $\mathcal{S}$ were in $E$ ). Hence, we have $r_{\mathcal{S}_{O}}(x) \leqslant r_{\mathcal{S}}(x)-1$, and thus

$$
\sum_{x \in O} 2^{-r_{\mathcal{S}}(x)} \leqslant \frac{1}{2} \sum_{x \in O} 2^{-r_{\mathcal{S}_{O}}(x)}
$$

Similarly, for every $x \in E$, we have

$$
\sum_{x \in E} 2^{-r_{\mathcal{S}}(x)} \leqslant \frac{1}{2} \sum_{x \in E} 2^{-r_{\mathcal{S}_{E}}(x)}
$$

We can then combine these to give

$$
\begin{aligned}
w(S) & =\sum_{x \in R} 2^{-r_{\mathcal{S}}(x)}+\sum_{x \in O} 2^{-r_{\mathcal{S}}(x)}+\sum_{x \in E} 2^{-r_{\mathcal{S}}(x)} \\
& \leqslant \frac{1}{2} \sum_{x \in R}\left(2^{-r_{S_{O}}(x)}+2^{-r_{\mathcal{S}_{E}}(x)}\right)+\frac{1}{2} \sum_{x \in O} 2^{-r_{\mathcal{S}_{O}}(x)}+\frac{1}{2} \sum_{x \in E} 2^{-r_{S_{E}}(x)} \\
& =\frac{1}{2}\left(\sum_{x \in \mathcal{S}_{O}} 2^{-r_{\mathcal{S}_{O}}(x)}+\sum_{x \in \mathcal{S}_{E}} 2^{-r \mathcal{S}_{E}(x)}\right) \quad\left(\text { since } \mathcal{S}_{O}=O \cup R \text { and } \mathcal{S}_{E}=E \cup R\right) \\
& \left.\left.=\frac{1}{2}\left(w\left(\mathcal{S}_{O}\right)+w\left(\mathcal{S}_{E}\right)\right)\right) \quad \text { (by definition of } w(\cdot)\right) \\
& \leqslant 1 \quad \text { (by the inductive hypothesis) }
\end{aligned}
$$

which completes the induction.
Comment 1. The sets $O$ and $E$ above are not the only ones we could have chosen. Indeed, we could instead have used the following definitions:

Let $d$ be the maximal scale between two distinct elements of $\mathcal{S}$; that is, $d=D\left(x_{1}, x_{n}\right)$. Let $O=\left\{x \in \mathcal{S}: D\left(x, x_{n}\right)=d\right\}$ (a 'left' part of the set) and let $E=\left\{x \in \mathcal{S}: D\left(x_{1}, x\right)=d\right\}$ (a 'right' part of the set). Note that these two sets are disjoint, and nonempty (since they contain $x_{1}$ and $x_{n}$ respectively). The rest of the proof is then the same as in Solution 1.

Comment 2. Another possible set $\mathcal{F}$ containing $2^{k}$ members could arise from considering a binary tree of height $k$, allocating a real number to each leaf, and trying to make the scale between the values of two leaves dependent only on the (graph) distance between them. The following construction makes this more precise.

We build up sets $\mathcal{F}_{k}$ recursively. Let $\mathcal{F}_{0}=\{0\}$, and then let $\mathcal{F}_{k+1}=\mathcal{F}_{k} \cup\left\{x+3 \cdot 4^{k}: x \in \mathcal{F}_{k}\right\}$ (i.e. each half of $\mathcal{F}_{k+1}$ is a copy of $\left.F_{k}\right)$. We have that $\mathcal{F}_{k}$ is contained in the interval $\left[0,4^{k+1}\right)$, and so it follows by induction on $k$ that every member of $F_{k+1}$ has $k$ different scales in its own half of $F_{k+1}$ (by the inductive hypothesis), and only the single scale $2 k+1$ in the other half of $F_{k+1}$.

Both of the constructions presented here have the property that every member of $\mathcal{F}$ has exactly $k$ different scales in $\mathcal{F}$. Indeed, it can be seen that this must hold (up to a slight perturbation) for any such maximal set. Suppose there were some element $x$ with only $k-1$ different scales in $\mathcal{F}$ (and every other element had at most $k$ different scales). Then we take some positive real $\epsilon$, and construct a new set $\mathcal{F}^{\prime}=\{y: y \in \mathcal{F}, y \leqslant x\} \cup\{y+\epsilon: y \in \mathcal{F}, y \geqslant x\}$. We have $\left|\mathcal{F}^{\prime}\right|=|\mathcal{F}|+1$, and if $\epsilon$ is sufficiently small then $\mathcal{F}^{\prime}$ will also satisfy the property that no member has more than $k$ different scales in $\mathcal{F}^{\prime}$.

This observation might be used to motivate the idea of weighting members of an arbitrary set $\mathcal{S}$ of reals according to how many different scales they have in $\mathcal{S}$.

## Geometry

G1. Let $A B C$ be a triangle. Circle $\Gamma$ passes through $A$, meets segments $A B$ and $A C$ again at points $D$ and $E$ respectively, and intersects segment $B C$ at $F$ and $G$ such that $F$ lies between $B$ and $G$. The tangent to circle $B D F$ at $F$ and the tangent to circle $C E G$ at $G$ meet at point $T$. Suppose that points $A$ and $T$ are distinct. Prove that line $A T$ is parallel to $B C$.
(Nigeria)
Solution. Notice that $\angle T F B=\angle F D A$ because $F T$ is tangent to circle $B D F$, and moreover $\angle F D A=\angle C G A$ because quadrilateral $A D F G$ is cyclic. Similarly, $\angle T G B=\angle G E C$ because $G T$ is tangent to circle $C E G$, and $\angle G E C=\angle C F A$. Hence,

$$
\begin{equation*}
\angle T F B=\angle C G A \quad \text { and } \quad \angle T G B=\angle C F A \tag{1}
\end{equation*}
$$



Triangles $F G A$ and $G F T$ have a common side $F G$, and by (1) their angles at $F, G$ are the same. So, these triangles are congruent. So, their altitudes starting from $A$ and $T$, respectively, are equal and hence $A T$ is parallel to line $B F G C$.

Comment. Alternatively, we can prove first that $T$ lies on $\Gamma$. For example, this can be done by showing that $\angle A F T=\angle A G T$ using (1). Then the statement follows as $\angle T A F=\angle T G F=\angle G F A$.

G2. Let $A B C$ be an acute-angled triangle and let $D, E$, and $F$ be the feet of altitudes from $A, B$, and $C$ to sides $B C, C A$, and $A B$, respectively. Denote by $\omega_{B}$ and $\omega_{C}$ the incircles of triangles $B D F$ and $C D E$, and let these circles be tangent to segments $D F$ and $D E$ at $M$ and $N$, respectively. Let line $M N$ meet circles $\omega_{B}$ and $\omega_{C}$ again at $P \neq M$ and $Q \neq N$, respectively. Prove that $M P=N Q$.
(Vietnam)
Solution. Denote the centres of $\omega_{B}$ and $\omega_{C}$ by $O_{B}$ and $O_{C}$, let their radii be $r_{B}$ and $r_{C}$, and let $B C$ be tangent to the two circles at $T$ and $U$, respectively.


From the cyclic quadrilaterals $A F D C$ and $A B D E$ we have

$$
\angle M D O_{B}=\frac{1}{2} \angle F D B=\frac{1}{2} \angle B A C=\frac{1}{2} \angle C D E=\angle O_{C} D N,
$$

so the right-angled triangles $D M O_{B}$ and $D N O_{C}$ are similar. The ratio of similarity between the two triangles is

$$
\frac{D N}{D M}=\frac{O_{C} N}{O_{B} M}=\frac{r_{C}}{r_{B}} .
$$

Let $\varphi=\angle D M N$ and $\psi=\angle M N D$. The lines $F M$ and $E N$ are tangent to $\omega_{B}$ and $\omega_{C}$, respectively, so

$$
\angle M T P=\angle F M P=\angle D M N=\varphi \quad \text { and } \quad \angle Q U N=\angle Q N E=\angle M N D=\psi
$$

(It is possible that $P$ or $Q$ coincides with $T$ or $U$, or lie inside triangles $D M T$ or $D U N$, respectively. To reduce case-sensitivity, we may use directed angles or simply ignore angles $M T P$ and $Q U N$.)

In the circles $\omega_{B}$ and $\omega_{C}$ the lengths of chords $M P$ and $N Q$ are

$$
M P=2 r_{B} \cdot \sin \angle M T P=2 r_{B} \cdot \sin \varphi \quad \text { and } \quad N Q=2 r_{C} \cdot \sin \angle Q U N=2 r_{C} \cdot \sin \psi
$$

By applying the sine rule to triangle $D N M$ we get

$$
\frac{D N}{D M}=\frac{\sin \angle D M N}{\sin \angle M N D}=\frac{\sin \varphi}{\sin \psi} .
$$

Finally, putting the above observations together, we get

$$
\frac{M P}{N Q}=\frac{2 r_{B} \sin \varphi}{2 r_{C} \sin \psi}=\frac{r_{B}}{r_{C}} \cdot \frac{\sin \varphi}{\sin \psi}=\frac{D M}{D N} \cdot \frac{\sin \varphi}{\sin \psi}=\frac{\sin \psi}{\sin \varphi} \cdot \frac{\sin \varphi}{\sin \psi}=1,
$$

so $M P=N Q$ as required.

G3. In triangle $A B C$, let $A_{1}$ and $B_{1}$ be two points on sides $B C$ and $A C$, and let $P$ and $Q$ be two points on segments $A A_{1}$ and $B B_{1}$, respectively, so that line $P Q$ is parallel to $A B$. On ray $P B_{1}$, beyond $B_{1}$, let $P_{1}$ be a point so that $\angle P P_{1} C=\angle B A C$. Similarly, on ray $Q A_{1}$, beyond $A_{1}$, let $Q_{1}$ be a point so that $\angle C Q_{1} Q=\angle C B A$. Show that points $P, Q, P_{1}$, and $Q_{1}$ are concyclic.
(Ukraine)
Solution 1. Throughout the solution we use oriented angles.
Let rays $A A_{1}$ and $B B_{1}$ intersect the circumcircle of $\triangle A C B$ at $A_{2}$ and $B_{2}$, respectively. By

$$
\angle Q P A_{2}=\angle B A A_{2}=\angle B B_{2} A_{2}=\angle Q B_{2} A_{2}
$$

points $P, Q, A_{2}, B_{2}$ are concyclic; denote the circle passing through these points by $\omega$. We shall prove that $P_{1}$ and $Q_{1}$ also lie on $\omega$.


By

$$
\angle C A_{2} A_{1}=\angle C A_{2} A=\angle C B A=\angle C Q_{1} Q=\angle C Q_{1} A_{1},
$$

points $C, Q_{1}, A_{2}, A_{1}$ are also concyclic. From that we get

$$
\angle Q Q_{1} A_{2}=\angle A_{1} Q_{1} A_{2}=\angle A_{1} C A_{2}=\angle B C A_{2}=\angle B A A_{2}=\angle Q P A_{2},
$$

so $Q_{1}$ lies on $\omega$.
It follows similarly that $P_{1}$ lies on $\omega$.
Solution 2. First consider the case when lines $P P_{1}$ and $Q Q_{1}$ intersect each other at some point $R$.

Let line $P Q$ meet the sides $A C$ and $B C$ at $E$ and $F$, respectively. Then

$$
\angle P P_{1} C=\angle B A C=\angle P E C,
$$

so points $C, E, P, P_{1}$ lie on a circle; denote that circle by $\omega_{P}$. It follows analogously that points $C, F, Q, Q_{1}$ lie on another circle; denote it by $\omega_{Q}$.

Let $A Q$ and $B P$ intersect at $T$. Applying Pappus' theorem to the lines $A A_{1} P$ and $B B_{1} Q$ provides that points $C=A B_{1} \cap B A_{1}, R=A_{1} Q \cap B_{1} P$ and $T=A Q \cap B P$ are collinear.

Let line $R C T$ meet $P Q$ and $A B$ at $S$ and $U$, respectively. From $A B \| P Q$ we obtain

$$
\frac{S P}{S Q}=\frac{U B}{U A}=\frac{S F}{S E},
$$

$$
S P \cdot S E=S Q \cdot S F
$$



So, point $S$ has equal powers with respect to $\omega_{P}$ and $\omega_{Q}$, hence line $R C S$ is their radical axis; then $R$ also has equal powers to the circles, so $R P \cdot R P_{1}=R Q \cdot R Q_{1}$, proving that points $P, P_{1}, Q, Q_{1}$ are indeed concyclic.

Now consider the case when $P P_{1}$ and $Q Q_{1}$ are parallel. Like in the previous case, let $A Q$ and $B P$ intersect at $T$. Applying Pappus' theorem again to the lines $A A_{1} P$ and $B B_{1} Q$, in this limit case it shows that line $C T$ is parallel to $P P_{1}$ and $Q Q_{1}$.

Let line $C T$ meet $P Q$ and $A B$ at $S$ and $U$, as before. The same calculation as in the previous case shows that $S P \cdot S E=S Q \cdot S F$, so $S$ lies on the radical axis between $\omega_{P}$ and $\omega_{Q}$.


Line $C S T$, that is the radical axis between $\omega_{P}$ and $\omega_{Q}$, is perpendicular to the line $\ell$ of centres of $\omega_{P}$ and $\omega_{Q}$. Hence, the chords $P P_{1}$ and $Q Q_{1}$ are perpendicular to $\ell$. So the quadrilateral $P P_{1} Q_{1} Q$ is an isosceles trapezium with symmetry axis $\ell$, and hence is cyclic.

Comment. There are several ways of solving the problem involving Pappus' theorem. For example, one may consider the points $K=P B_{1} \cap B C$ and $L=Q A_{1} \cap A C$. Applying Pappus' theorem to the lines $A A_{1} P$ and $Q B_{1} B$ we get that $K, L$, and $P Q \cap A B$ are collinear, i.e. that $K L \| A B$. Therefore, cyclicity of $P, Q, P_{1}$, and $Q_{1}$ is equivalent to that of $K, L, P_{1}$, and $Q_{1}$. The latter is easy after noticing that $C$ also lies on that circle. Indeed, e.g. $\angle(L K, L C)=\angle(A B, A C)=\angle\left(P_{1} K, P_{1} C\right)$ shows that $K$ lies on circle $K L C$.

This approach also has some possible degeneracy, as the points $K$ and $L$ may happen to be ideal.

G4. Let $P$ be a point inside triangle $A B C$. Let $A P$ meet $B C$ at $A_{1}$, let $B P$ meet $C A$ at $B_{1}$, and let $C P$ meet $A B$ at $C_{1}$. Let $A_{2}$ be the point such that $A_{1}$ is the midpoint of $P A_{2}$, let $B_{2}$ be the point such that $B_{1}$ is the midpoint of $P B_{2}$, and let $C_{2}$ be the point such that $C_{1}$ is the midpoint of $P C_{2}$. Prove that points $A_{2}, B_{2}$, and $C_{2}$ cannot all lie strictly inside the circumcircle of triangle $A B C$.
(Australia)


Solution 1. Since

$$
\angle A P B+\angle B P C+\angle C P A=2 \pi=(\pi-\angle A C B)+(\pi-\angle B A C)+(\pi-\angle C B A),
$$

at least one of the following inequalities holds:

$$
\angle A P B \geqslant \pi-\angle A C B, \quad \angle B P C \geqslant \pi-\angle B A C, \quad \angle C P A \geqslant \pi-\angle C B A .
$$

Without loss of generality, we assume that $\angle B P C \geqslant \pi-\angle B A C$. We have $\angle B P C>\angle B A C$ because $P$ is inside $\triangle A B C$. So $\angle B P C \geqslant \max (\angle B A C, \pi-\angle B A C)$ and hence

$$
\begin{equation*}
\sin \angle B P C \leqslant \sin \angle B A C . \tag{*}
\end{equation*}
$$

Let the rays $A P, B P$, and $C P$ cross the circumcircle $\Omega$ again at $A_{3}, B_{3}$, and $C_{3}$, respectively. We will prove that at least one of the ratios $\frac{P B_{1}}{B_{1} B_{3}}$ and $\frac{P C_{1}}{C_{1} C_{3}}$ is at least 1 , which yields that one of the points $B_{2}$ and $C_{2}$ does not lie strictly inside $\Omega$.

Because $A, B, C, B_{3}$ lie on a circle, the triangles $C B_{1} B_{3}$ and $B B_{1} A$ are similar, so

$$
\frac{C B_{1}}{B_{1} B_{3}}=\frac{B B_{1}}{B_{1} A} .
$$

Applying the sine rule we obtain

$$
\frac{P B_{1}}{B_{1} B_{3}}=\frac{P B_{1}}{C B_{1}} \cdot \frac{C B_{1}}{B_{1} B_{3}}=\frac{P B_{1}}{C B_{1}} \cdot \frac{B B_{1}}{B_{1} A}=\frac{\sin \angle A C P}{\sin \angle B P C} \cdot \frac{\sin \angle B A C}{\sin \angle P B A} .
$$

Similarly,

$$
\frac{P C_{1}}{C_{1} C_{3}}=\frac{\sin \angle P B A}{\sin \angle B P C} \cdot \frac{\sin \angle B A C}{\sin \angle A C P} .
$$

Multiplying these two equations we get

$$
\frac{P B_{1}}{B_{1} B_{3}} \cdot \frac{P C_{1}}{C_{1} C_{3}}=\frac{\sin ^{2} \angle B A C}{\sin ^{2} \angle B P C} \geqslant 1
$$

using (*), which yields the desired conclusion.

Comment. It also cannot happen that all three points $A_{2}, B_{2}$, and $C_{2}$ lie strictly outside $\Omega$. The same proof works almost literally, starting by assuming without loss of generality that $\angle B P C \leqslant \pi-\angle B A C$ and using $\angle B P C>\angle B A C$ to deduce that $\sin \angle B P C \geqslant \sin \angle B A C$. It is possible for $A_{2}, B_{2}$, and $C_{2}$ all to lie on the circumcircle; from the above solution we may derive that this happens if and only if $P$ is the orthocentre of the triangle $A B C$, (which lies strictly inside $A B C$ if and only if $A B C$ is acute).

Solution 2. Define points $A_{3}, B_{3}$, and $C_{3}$ as in Solution 1. Assume for the sake of contradiction that $A_{2}, B_{2}$, and $C_{2}$ all lie strictly inside circle $A B C$. It follows that $P A_{1}<A_{1} A_{3}, P B_{1}<B_{1} B_{3}$, and $P C_{1}<C_{1} C_{3}$.

Observe that $\triangle P B C_{3} \sim \triangle P C B_{3}$. Let $X$ be the point on side $P B_{3}$ that corresponds to point $C_{1}$ on side $P C_{3}$ under this similarity. In other words, $X$ lies on segment $P B_{3}$ and satisfies $P X: X B_{3}=P C_{1}: C_{1} C_{3}$. It follows that

$$
\angle X C P=\angle P B C_{1}=\angle B_{3} B A=\angle B_{3} C B_{1} .
$$

Hence lines $C X$ and $C B_{1}$ are isogonal conjugates in $\triangle P C B_{3}$.


Let $Y$ be the foot of the bisector of $\angle B_{3} C P$ in $\triangle P C B_{3}$. Since $P C_{1}<C_{1} C_{3}$, we have $P X<X B_{3}$. Also, we have $P Y<Y B_{3}$ because $P B_{1}<B_{1} B_{3}$ and $Y$ lies between $X$ and $B_{1}$. By the angle bisector theorem in $\triangle P C B_{3}$, we have $P Y: Y B_{3}=P C: C B_{3}$. So $P C<C B_{3}$ and it follows that $\angle P B_{3} C<\angle C P B_{3}$. Now since $\angle P B_{3} C=\angle B B_{3} C=\angle B A C$, we have

$$
\angle B A C<\angle C P B_{3} .
$$

Similarly, we have

$$
\angle C B A<\angle A P C_{3} \quad \text { and } \quad \angle A C B<\angle B P A_{3}=\angle B_{3} P A .
$$

Adding these three inequalities yields $\pi<\pi$, and this contradiction concludes the proof.

Solution 3. Choose coordinates such that the circumcentre of $\triangle A B C$ is at the origin and the circumradius is 1 . Then we may think of $A, B$, and $C$ as vectors in $\mathbb{R}^{2}$ such that

$$
|A|^{2}=|B|^{2}=|C|^{2}=1
$$

$P$ may be represented as a convex combination $\alpha A+\beta B+\gamma C$ where $\alpha, \beta, \gamma>0$ and $\alpha+\beta+\gamma=1$. Then

$$
A_{1}=\frac{\beta B+\gamma C}{\beta+\gamma}=\frac{1}{1-\alpha} P-\frac{\alpha}{1-\alpha} A,
$$

so

$$
A_{2}=2 A_{1}-P=\frac{1+\alpha}{1-\alpha} P-\frac{2 \alpha}{1-\alpha} A
$$

Hence

$$
\left|A_{2}\right|^{2}=\left(\frac{1+\alpha}{1-\alpha}\right)^{2}|P|^{2}+\left(\frac{2 \alpha}{1-\alpha}\right)^{2}|A|^{2}-\frac{4 \alpha(1+\alpha)}{(1-\alpha)^{2}} A \cdot P .
$$

Using $|A|^{2}=1$ we obtain

$$
\begin{equation*}
\frac{(1-\alpha)^{2}}{2(1+\alpha)}\left|A_{2}\right|^{2}=\frac{1+\alpha}{2}|P|^{2}+\frac{2 \alpha^{2}}{1+\alpha}-2 \alpha A \cdot P . \tag{1}
\end{equation*}
$$

Likewise

$$
\begin{equation*}
\frac{(1-\beta)^{2}}{2(1+\beta)}\left|B_{2}\right|^{2}=\frac{1+\beta}{2}|P|^{2}+\frac{2 \beta^{2}}{1+\beta}-2 \beta B \cdot P \tag{2}
\end{equation*}
$$

and

$$
\begin{equation*}
\frac{(1-\gamma)^{2}}{2(1+\gamma)}\left|C_{2}\right|^{2}=\frac{1+\gamma}{2}|P|^{2}+\frac{2 \gamma^{2}}{1+\gamma}-2 \gamma C \cdot P \tag{3}
\end{equation*}
$$

Summing (1), (2) and (3) we obtain on the LHS the positive linear combination

$$
\text { LHS }=\frac{(1-\alpha)^{2}}{2(1+\alpha)}\left|A_{2}\right|^{2}+\frac{(1-\beta)^{2}}{2(1+\beta)}\left|B_{2}\right|^{2}+\frac{(1-\gamma)^{2}}{2(1+\gamma)}\left|C_{2}\right|^{2}
$$

and on the RHS the quantity

$$
\left(\frac{1+\alpha}{2}+\frac{1+\beta}{2}+\frac{1+\gamma}{2}\right)|P|^{2}+\left(\frac{2 \alpha^{2}}{1+\alpha}+\frac{2 \beta^{2}}{1+\beta}+\frac{2 \gamma^{2}}{1+\gamma}\right)-2(\alpha A \cdot P+\beta B \cdot P+\gamma C \cdot P) .
$$

The first term is $2|P|^{2}$ and the last term is $-2 P \cdot P$, so

$$
\begin{aligned}
\mathrm{RHS} & =\left(\frac{2 \alpha^{2}}{1+\alpha}+\frac{2 \beta^{2}}{1+\beta}+\frac{2 \gamma^{2}}{1+\gamma}\right) \\
& =\frac{3 \alpha-1}{2}+\frac{(1-\alpha)^{2}}{2(1+\alpha)}+\frac{3 \beta-1}{2}+\frac{(1-\beta)^{2}}{2(1+\beta)}+\frac{3 \gamma-1}{2}+\frac{(1-\gamma)^{2}}{2(1+\gamma)} \\
& =\frac{(1-\alpha)^{2}}{2(1+\alpha)}+\frac{(1-\beta)^{2}}{2(1+\beta)}+\frac{(1-\gamma)^{2}}{2(1+\gamma)} .
\end{aligned}
$$

Here we used the fact that

$$
\frac{3 \alpha-1}{2}+\frac{3 \beta-1}{2}+\frac{3 \gamma-1}{2}=0 .
$$

We have shown that a linear combination of $\left|A_{1}\right|^{2},\left|B_{1}\right|^{2}$, and $\left|C_{1}\right|^{2}$ with positive coefficients is equal to the sum of the coefficients. Therefore at least one of $\left|A_{1}\right|^{2},\left|B_{1}\right|^{2}$, and $\left|C_{1}\right|^{2}$ must be at least 1 , as required.

Comment. This proof also works when $P$ is any point for which $\alpha, \beta, \gamma>-1, \alpha+\beta+\gamma=1$, and $\alpha, \beta, \gamma \neq 1$. (In any cases where $\alpha=1$ or $\beta=1$ or $\gamma=1$, some points in the construction are not defined.)

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G5. Let $A B C D E$ be a convex pentagon with $C D=D E$ and $\angle E D C \neq 2 \cdot \angle A D B$. Suppose that a point $P$ is located in the interior of the pentagon such that $A P=A E$ and $B P=B C$. Prove that $P$ lies on the diagonal $C E$ if and only if area $(B C D)+\operatorname{area}(A D E)=$ $\operatorname{area}(A B D)+\operatorname{area}(A B P)$.
(Hungary)
Solution 1. Let $P^{\prime}$ be the reflection of $P$ across line $A B$, and let $M$ and $N$ be the midpoints of $P^{\prime} E$ and $P^{\prime} C$ respectively. Convexity ensures that $P^{\prime}$ is distinct from both $E$ and $C$, and hence from both $M$ and $N$. We claim that both the area condition and the collinearity condition in the problem are equivalent to the condition that the (possibly degenerate) right-angled triangles $A P^{\prime} M$ and $B P^{\prime} N$ are directly similar (equivalently, $A P^{\prime} E$ and $B P^{\prime} C$ are directly similar).


For the equivalence with the collinearity condition, let $F$ denote the foot of the perpendicular from $P^{\prime}$ to $A B$, so that $F$ is the midpoint of $P P^{\prime}$. We have that $P$ lies on $C E$ if and only if $F$ lies on $M N$, which occurs if and only if we have the equality $\angle A F M=\angle B F N$ of signed angles modulo $\pi$. By concyclicity of $A P^{\prime} F M$ and $B F P^{\prime} N$, this is equivalent to $\angle A P^{\prime} M=\angle B P^{\prime} N$, which occurs if and only if $A P^{\prime} M$ and $B P^{\prime} N$ are directly similar.


For the other equivalence with the area condition, we have the equality of signed areas $\operatorname{area}(A B D)+\operatorname{area}(A B P)=\operatorname{area}\left(A P^{\prime} B D\right)=\operatorname{area}\left(A P^{\prime} D\right)+\operatorname{area}\left(B D P^{\prime}\right)$. Using the identity area $(A D E)-\operatorname{area}\left(A P^{\prime} D\right)=\operatorname{area}(A D E)+\operatorname{area}\left(A D P^{\prime}\right)=2$ area $(A D M)$, and similarly for $B$, we find that the area condition is equivalent to the equality

$$
\operatorname{area}(D A M)=\operatorname{area}(D B N)
$$

Now note that $A$ and $B$ lie on the perpendicular bisectors of $P^{\prime} E$ and $P^{\prime} C$, respectively. If we write $G$ and $H$ for the feet of the perpendiculars from $D$ to these perpendicular bisectors respectively, then this area condition can be rewritten as

$$
M A \cdot G D=N B \cdot H D
$$

(In this condition, we interpret all lengths as signed lengths according to suitable conventions: for instance, we orient $P^{\prime} E$ from $P^{\prime}$ to $E$, orient the parallel line $D H$ in the same direction, and orient the perpendicular bisector of $P^{\prime} E$ at an angle $\pi / 2$ clockwise from the oriented segment $P^{\prime} E$ - we adopt the analogous conventions at $B$.)


To relate the signed lengths $G D$ and $H D$ to the triangles $A P^{\prime} M$ and $B P^{\prime} N$, we use the following calculation.
Claim. Let $\Gamma$ denote the circle centred on $D$ with both $E$ and $C$ on the circumference, and $h$ the power of $P^{\prime}$ with respect to $\Gamma$. Then we have the equality

$$
G D \cdot P^{\prime} M=H D \cdot P^{\prime} N=\frac{1}{4} h \neq 0 .
$$

Proof. Firstly, we have $h \neq 0$, since otherwise $P^{\prime}$ would lie on $\Gamma$, and hence the internal angle bisectors of $\angle E D P^{\prime}$ and $\angle P^{\prime} D C$ would pass through $A$ and $B$ respectively. This would violate the angle inequality $\angle E D C \neq 2 \cdot \angle A D B$ given in the question.

Next, let $E^{\prime}$ denote the second point of intersection of $P^{\prime} E$ with $\Gamma$, and let $E^{\prime \prime}$ denote the point on $\Gamma$ diametrically opposite $E^{\prime}$, so that $E^{\prime \prime} E$ is perpendicular to $P^{\prime} E$. The point $G$ lies on the perpendicular bisectors of the sides $P^{\prime} E$ and $E E^{\prime \prime}$ of the right-angled triangle $P^{\prime} E E^{\prime \prime}$; it follows that $G$ is the midpoint of $P^{\prime} E^{\prime \prime}$. Since $D$ is the midpoint of $E^{\prime} E^{\prime \prime}$, we have that $G D=\frac{1}{2} P^{\prime} E^{\prime}$. Since $P^{\prime} M=\frac{1}{2} P^{\prime} E$, we have $G D \cdot P^{\prime} M=\frac{1}{4} P^{\prime} E^{\prime} \cdot P^{\prime} E=\frac{1}{4} h$. The other equality $H D \cdot P^{\prime} N$ follows by exactly the same argument.


From this claim, we see that the area condition is equivalent to the equality

$$
\left(M A: P^{\prime} M\right)=\left(N B: P^{\prime} N\right)
$$

of ratios of signed lengths, which is equivalent to direct similarity of $A P^{\prime} M$ and $B P^{\prime} N$, as desired.

Solution 2. Along the perpendicular bisector of $C E$, define the linear function

$$
f(X)=\operatorname{area}(B C X)+\operatorname{area}(A X E)-\operatorname{area}(A B X)-\operatorname{area}(A B P),
$$

where, from now on, we always use signed areas. Thus, we want to show that $C, P, E$ are collinear if and only if $f(D)=0$.


Let $P^{\prime}$ be the reflection of $P$ across line $A B$. The point $P^{\prime}$ does not lie on the line $C E$. To see this, we let $A^{\prime \prime}$ and $B^{\prime \prime}$ be the points obtained from $A$ and $B$ by dilating with scale factor 2 about $P^{\prime}$, so that $P$ is the orthogonal projection of $P^{\prime}$ onto $A^{\prime \prime} B^{\prime \prime}$. Since $A$ lies on the perpendicular bisector of $P^{\prime} E$, the triangle $A^{\prime \prime} E P^{\prime}$ is right-angled at $E$ (and $B^{\prime \prime} C P^{\prime}$ similarly). If $P^{\prime}$ were to lie on $C E$, then the lines $A^{\prime \prime} E$ and $B^{\prime \prime} C$ would be perpendicular to $C E$ and $A^{\prime \prime}$ and $B^{\prime \prime}$ would lie on the opposite side of $C E$ to $D$. It follows that the line $A^{\prime \prime} B^{\prime \prime}$ does not meet triangle $C D E$, and hence point $P$ does not lie inside $C D E$. But then $P$ must lie inside $A B C E$, and it is clear that such a point cannot reflect to a point $P^{\prime}$ on $C E$.

We thus let $O$ be the centre of the circle $C E P^{\prime}$. The lines $A O$ and $B O$ are the perpendicular bisectors of $E P^{\prime}$ and $C P^{\prime}$, respectively, so

$$
\begin{aligned}
\operatorname{area}(B C O)+\operatorname{area}(A O E) & =\operatorname{area}\left(O P^{\prime} B\right)+\operatorname{area}\left(P^{\prime} O A\right)=\operatorname{area}\left(P^{\prime} B O A\right) \\
& =\operatorname{area}(A B O)+\operatorname{area}\left(B A P^{\prime}\right)=\operatorname{area}(A B O)+\operatorname{area}(A B P),
\end{aligned}
$$

and hence $f(O)=0$.
Notice that if point $O$ coincides with $D$ then points $A, B$ lie in angle domain $C D E$ and $\angle E O C=2 \cdot \angle A O B$, which is not allowed. So, $O$ and $D$ must be distinct. Since $f$ is linear and vanishes at $O$, it follows that $f(D)=0$ if and only if $f$ is constant zero - we want to show this occurs if and only if $C, P, E$ are collinear.


In the one direction, suppose firstly that $C, P, E$ are not collinear, and let $T$ be the centre of the circle $C E P$. The same calculation as above provides

$$
\operatorname{area}(B C T)+\operatorname{area}(A T E)=\operatorname{area}(P B T A)=\operatorname{area}(A B T)-\operatorname{area}(A B P)
$$

$$
f(T)=-2 \operatorname{area}(A B P) \neq 0
$$

Hence, the linear function $f$ is nonconstant with its zero is at $O$, so that $f(D) \neq 0$.
In the other direction, suppose that the points $C, P, E$ are collinear. We will show that $f$ is constant zero by finding a second point (other than $O$ ) at which it vanishes.


Let $Q$ be the reflection of $P$ across the midpoint of $A B$, so $P A Q B$ is a parallelogram. It is easy to see that $Q$ is on the perpendicular bisector of $C E$; for instance if $A^{\prime}$ and $B^{\prime}$ are the points produced from $A$ and $B$ by dilating about $P$ with scale factor 2, then the projection of $Q$ to $C E$ is the midpoint of the projections of $A^{\prime}$ and $B^{\prime}$, which are $E$ and $C$ respectively. The triangles $B C Q$ and $A Q E$ are indirectly congruent, so

$$
f(Q)=(\operatorname{area}(B C Q)+\operatorname{area}(A Q E))-(\operatorname{area}(A B Q)-\operatorname{area}(B A P))=0-0=0
$$

The points $O$ and $Q$ are distinct. To see this, consider the circle $\omega$ centred on $Q$ with $P^{\prime}$ on the circumference; since triangle $P P^{\prime} Q$ is right-angled at $P^{\prime}$, it follows that $P$ lies outside $\omega$. On the other hand, $P$ lies between $C$ and $E$ on the line $C P E$. It follows that $C$ and $E$ cannot both lie on $\omega$, so that $\omega$ is not the circle $C E P^{\prime}$ and $Q \neq O$.

Since $O$ and $Q$ are distinct zeroes of the linear function $f$, we have $f(D)=0$ as desired.
Comment 1. The condition $\angle E D C \neq 2 \cdot \angle A D B$ cannot be omitted. If $D$ is the centre of circle $C E P^{\prime}$, then the condition on triangle areas is satisfied automatically, without having $P$ on line $C E$.

Comment 2. The "only if" part of this problem is easier than the "if" part. For example, in the second part of Solution 2, the triangles $E A Q$ and $Q B C$ are indirectly congruent, so the sum of their areas is 0 , and $D C Q E$ is a kite. Now one can easily see that $\angle(A Q, D E)=\angle(C D, C B)$ and $\angle(B Q, D C)=\angle(E D, E A)$, whence $\operatorname{area}(B C D)=\operatorname{area}(A Q D)+\operatorname{area}(E Q A)$ and area $(A D E)=$ $\operatorname{area}(B D Q)+\operatorname{area}(B Q C)$, which yields the result.

Comment 3. The origin of the problem is the following observation. Let $A B D H$ be a tetrahedron and consider the sphere $\mathcal{S}$ that is tangent to the four face planes, internally to planes $A D H$ and $B D H$ and externally to $A B D$ and $A B H$ (or vice versa). It is known that the sphere $\mathcal{S}$ exists if and only if area $(A D H)+\operatorname{area}(B D H) \neq \operatorname{area}(A B H)+\operatorname{area}(A B D)$; this relation comes from the usual formula for the volume of the tetrahedron.

Let $T, T_{a}, T_{b}, T_{d}$ be the points of tangency between the sphere and the four planes, as shown in the picture. Rotate the triangle $A B H$ inward, the triangles $B D H$ and $A D H$ outward, into the triangles $A B P, B D C$ and $A D E$, respectively, in the plane $A B D$. Notice that the points $T_{d}, T_{a}, T_{b}$ are rotated to $T$, so we have $H T_{a}=H T_{b}=H T_{d}=P T=C T=E T$. Therefore, the point $T$ is the centre of the circle $C E P$. Hence, if the sphere exists then $C, E, P$ cannot be collinear.

If the condition $\angle E D C \neq 2 \cdot \angle A D B$ is replaced by the constraint that the angles $\angle E D A, \angle A D B$ and $\angle B D C$ satisfy the triangle inequality, it enables reconstructing the argument with the tetrahedron and the tangent sphere.


G6. Let $I$ be the incentre of acute-angled triangle $A B C$. Let the incircle meet $B C, C A$, and $A B$ at $D, E$, and $F$, respectively. Let line $E F$ intersect the circumcircle of the triangle at $P$ and $Q$, such that $F$ lies between $E$ and $P$. Prove that $\angle D P A+\angle A Q D=\angle Q I P$.
(Slovakia)
Solution 1. Let $N$ and $M$ be the midpoints of the arcs $\widehat{B C}$ of the circumcircle, containing and opposite vertex $A$, respectively. By $\angle F A E=\angle B A C=\angle B N C$, the right-angled kites $A F I E$ and $N B M C$ are similar. Consider the spiral similarity $\varphi$ (dilation in case of $A B=A C$ ) that moves AFIE to $N B M C$. The directed angle in which $\varphi$ changes directions is $\angle(A F, N B)$, same as $\angle(A P, N P)$ and $\angle(A Q, N Q)$; so lines $A P$ and $A Q$ are mapped to lines $N P$ and $N Q$, respectively. Line $E F$ is mapped to $B C$; we can see that the intersection points $P=E F \cap A P$ and $Q=E F \cap A Q$ are mapped to points $B C \cap N P$ and $B C \cap N Q$, respectively. Denote these points by $P^{\prime}$ and $Q^{\prime}$, respectively.


Let $L$ be the midpoint of $B C$. We claim that points $P, Q, D, L$ are concyclic (if $D=L$ then line $B C$ is tangent to circle $P Q D$ ). Let $P Q$ and $B C$ meet at $Z$. By applying Menelaus' theorem to triangle $A B C$ and line $E F Z$, we have

$$
\frac{B D}{D C}=\frac{B F}{F A} \cdot \frac{A E}{E C}=-\frac{B Z}{Z C},
$$

so the pairs $B, C$ and $D, Z$ are harmonic. It is well-known that this implies $Z B \cdot Z C=Z D \cdot Z L$. (The inversion with pole $Z$ that swaps $B$ and $C$ sends $Z$ to infinity and $D$ to the midpoint of $B C$, because the cross-ratio is preserved.) Hence, $Z D \cdot Z L=Z B \cdot Z C=Z P \cdot Z Q$ by the power of $Z$ with respect to the circumcircle; this proves our claim.

By $\angle M P P^{\prime}=\angle M Q Q^{\prime}=\angle M L P^{\prime}=\angle M L Q^{\prime}=90^{\circ}$, the quadrilaterals $M L P P^{\prime}$ and $M L Q Q^{\prime}$ are cyclic. Then the problem statement follows by

$$
\begin{aligned}
\angle D P A+\angle A Q D & =360^{\circ}-\angle P A Q-\angle Q D P=360^{\circ}-\angle P N Q-\angle Q L P \\
& =\angle L P N+\angle N Q L=\angle P^{\prime} M L+\angle L M Q^{\prime}=\angle P^{\prime} M Q^{\prime}=\angle P I Q .
\end{aligned}
$$

Solution 2. Define the point $M$ and the same spiral similarity $\varphi$ as in the previous solution. (The point $N$ is not necessary.) It is well-known that the centre of the spiral similarity that maps $F, E$ to $B, C$ is the Miquel point of the lines $F E, B C, B F$ and $C E$; that is, the second intersection of circles $A B C$ and $A E F$. Denote that point by $S$.

By $\varphi(F)=B$ and $\varphi(E)=C$ the triangles $S B F$ and $S C E$ are similar, so we have

$$
\frac{S B}{S C}=\frac{B F}{C E}=\frac{B D}{C D}
$$

By the converse of the angle bisector theorem, that indicates that line $S D$ bisects $\angle B S C$ and hence passes through $M$.

Let $K$ be the intersection point of lines $E F$ and $S I$. Notice that $\varphi$ sends points $S, F, E, I$ to $S, B, C, M$, so $\varphi(K)=\varphi(F E \cap S I)=B C \cap S M=D$. By $\varphi(I)=M$, we have $K D \| I M$.


We claim that triangles $S P I$ and $S D Q$ are similar, and so are triangles $S P D$ and $S I Q$. Let ray $S I$ meet the circumcircle again at $L$. Note that the segment $E F$ is perpendicular to the angle bisector $A M$. Then by $\angle A M L=\angle A S L=\angle A S I=90^{\circ}$, we have $M L \| P Q$. Hence, $\widetilde{P L}=\widetilde{M Q}$ and therefore $\angle P S L=\angle M S Q=\angle D S Q$. By $\angle Q P S=\angle Q M S$, the triangles $S P K$ and $S M Q$ are similar. Finally,

$$
\frac{S P}{S I}=\frac{S P}{S K} \cdot \frac{S K}{S I}=\frac{S M}{S Q} \cdot \frac{S D}{S M}=\frac{S D}{S Q}
$$

shows that triangles $S P I$ and $S D Q$ are similar. The second part of the claim can be proved analogously.

Now the problem statement can be proved by

$$
\angle D P A+\angle A Q D=\angle D P S+\angle S Q D=\angle Q I S+\angle S I P=\angle Q I P
$$

Solution 3. Denote the circumcircle of triangle $A B C$ by $\Gamma$, and let rays $P D$ and $Q D$ meet $\Gamma$ again at $V$ and $U$, respectively. We will show that $A U \perp I P$ and $A V \perp I Q$. Then the problem statement will follow as

$$
\angle D P A+\angle A Q D=\angle V U A+\angle A V U=180^{\circ}-\angle U A V=\angle Q I P .
$$

Let $M$ be the midpoint of arc $\widehat{B U V C}$ and let $N$ be the midpoint of $\operatorname{arc} \widehat{C A B}$; the lines $A I M$ and $A N$ being the internal and external bisectors of angle $B A C$, respectively, are perpendicular. Let the tangents drawn to $\Gamma$ at $B$ and $C$ meet at $R$; let line $P Q$ meet $A U, A I, A V$ and $B C$ at $X, T, Y$ and $Z$, respectively.

As in Solution 1, we observe that the pairs $B, C$ and $D, Z$ are harmonic. Projecting these points from $Q$ onto the circumcircle, we can see that $B, C$ and $U, P$ are also harmonic. Analogously, the pair $V, Q$ is harmonic with $B, C$. Consider the inversion about the circle with centre $R$, passing through $B$ and $C$. Points $B$ and $C$ are fixed points, so this inversion exchanges every point of $\Gamma$ by its harmonic pair with respect to $B, C$. In particular, the inversion maps points $B, C, N, U, V$ to points $B, C, M, P, Q$, respectively.

Combine the inversion with projecting $\Gamma$ from $A$ to line $P Q$; the points $B, C, M, P, Q$ are projected to $F, E, T, P, Q$, respectively.


The combination of these two transformations is projective map from the lines $A B, A C$, $A N, A U, A V$ to $I F, I E, I T, I P, I Q$, respectively. On the other hand, we have $A B \perp I F$, $A C \perp I E$ and $A N \perp A T$, so the corresponding lines in these two pencils are perpendicular. This proves $A U \perp I P$ and $A V \perp I Q$, and hence completes the solution.

G7. The incircle $\omega$ of acute-angled scalene triangle $A B C$ has centre $I$ and meets sides $B C$, $C A$, and $A B$ at $D, E$, and $F$, respectively. The line through $D$ perpendicular to $E F$ meets $\omega$ again at $R$. Line $A R$ meets $\omega$ again at $P$. The circumcircles of triangles $P C E$ and $P B F$ meet again at $Q \neq P$. Prove that lines $D I$ and $P Q$ meet on the external bisector of angle $B A C$.
(India)
Common remarks. Throughout the solution, $\angle(a, b)$ denotes the directed angle between lines $a$ and $b$, measured modulo $\pi$.

## Solution 1.

Step 1. The external bisector of $\angle B A C$ is the line through $A$ perpendicular to $I A$. Let $D I$ meet this line at $L$ and let $D I$ meet $\omega$ at $K$. Let $N$ be the midpoint of $E F$, which lies on $I A$ and is the pole of line $A L$ with respect to $\omega$. Since $A N \cdot A I=A E^{2}=A R \cdot A P$, the points $R$, $N, I$, and $P$ are concyclic. As $I R=I P$, the line $N I$ is the external bisector of $\angle P N R$, so $P N$ meets $\omega$ again at the point symmetric to $R$ with respect to $A N$ - i.e. at $K$.

Let $D N$ cross $\omega$ again at $S$. Opposite sides of any quadrilateral inscribed in the circle $\omega$ meet on the polar line of the intersection of the diagonals with respect to $\omega$. Since $L$ lies on the polar line $A L$ of $N$ with respect to $\omega$, the line $P S$ must pass through $L$. Thus it suffices to prove that the points $S, Q$, and $P$ are collinear.


Step 2. Let $\Gamma$ be the circumcircle of $\triangle B I C$. Notice that

$$
\begin{aligned}
& \angle(B Q, Q C)=\angle(B Q, Q P)+\angle(P Q, Q C)=\angle(B F, F P)+\angle(P E, E C) \\
&=\angle(E F, E P)+\angle(F P, F E)=\angle(F P, E P)=\angle(D F, D E)=\angle(B I, I C),
\end{aligned}
$$

so $Q$ lies on $\Gamma$. Let $Q P$ meet $\Gamma$ again at $T$. It will now suffice to prove that $S, P$, and $T$ are collinear. Notice that $\angle(B I, I T)=\angle(B Q, Q T)=\angle(B F, F P)=\angle(F K, K P)$. Note $F D \perp F K$ and $F D \perp B I$ so $F K \| B I$ and hence $I T$ is parallel to the line $K N P$. Since $D I=I K$, the line $I T$ crosses $D N$ at its midpoint $M$.
Step 3. Let $F^{\prime}$ and $E^{\prime}$ be the midpoints of $D E$ and $D F$, respectively. Since $D E^{\prime} \cdot E^{\prime} F=D E^{\prime 2}=$ $B E^{\prime} \cdot E^{\prime} I$, the point $E^{\prime}$ lies on the radical axis of $\omega$ and $\Gamma$; the same holds for $F^{\prime}$. Therefore, this radical axis is $E^{\prime} F^{\prime}$, and it passes through $M$. Thus $I M \cdot M T=D M \cdot M S$, so $S, I, D$, and $T$ are concyclic. This shows $\angle(D S, S T)=\angle(D I, I T)=\angle(D K, K P)=\angle(D S, S P)$, whence the points $S, P$, and $T$ are collinear, as desired.


Comment. Here is a longer alternative proof in step 1 that $P, S$, and $L$ are collinear, using a circular inversion instead of the fact that opposite sides of a quadrilateral inscribed in a circle $\omega$ meet on the polar line with respect to $\omega$ of the intersection of the diagonals. Let $G$ be the foot of the altitude from $N$ to the line DIKL. Observe that $N, G, K, S$ are concyclic (opposite right angles) so

$$
\angle D I P=2 \angle D K P=\angle G K N+\angle D S P=\angle G S N+\angle N S P=\angle G S P,
$$

hence $I, G, S, P$ are concyclic. We have $I G \cdot I L=I N \cdot I A=r^{2}$ since $\triangle I G N \sim \triangle I A L$. Inverting the circle $I G S P$ in circle $\omega$, points $P$ and $S$ are fixed and $G$ is taken to $L$ so we find that $P, S$, and $L$ are collinear.

Solution 2. We start as in Solution 1. Namely, we introduce the same points $K, L, N$, and $S$, and show that the triples $(P, N, K)$ and $(P, S, L)$ are collinear. We conclude that $K$ and $R$ are symmetric in $A I$, and reduce the problem statement to showing that $P, Q$, and $S$ are collinear.

Step 1. Let $A R$ meet the circumcircle $\Omega$ of $A B C$ again at $X$. The lines $A R$ and $A K$ are isogonal in the angle $B A C$; it is well known that in this case $X$ is the tangency point of $\Omega$ with the $A$-mixtilinear circle. It is also well known that for this point $X$, the line $X I$ crosses $\Omega$ again at the midpoint $M^{\prime}$ of arc $B A C$.

Step 2. Denote the circles $B F P$ and $C E P$ by $\Omega_{B}$ and $\Omega_{C}$, respectively. Let $\Omega_{B}$ cross $A R$ and $E F$ again at $U$ and $Y$, respectively. We have

$$
\angle(U B, B F)=\angle(U P, P F)=\angle(R P, P F)=\angle(R F, F A),
$$

so $U B \| R F$.


Next, we show that the points $B, I, U$, and $X$ are concyclic. Since

$$
\angle(U B, U X)=\angle(R F, R X)=\angle(A F, A R)+\angle(F R, F A)=\angle\left(M^{\prime} B, M^{\prime} X\right)+\angle(D R, D F)
$$

it suffices to prove $\angle(I B, I X)=\angle\left(M^{\prime} B, M^{\prime} X\right)+\angle(D R, D F)$, or $\angle\left(I B, M^{\prime} B\right)=\angle(D R, D F)$. But both angles equal $\angle(C I, C B)$, as desired. (This is where we used the fact that $M^{\prime}$ is the midpoint of $\operatorname{arc} B A C$ of $\Omega$.)

It follows now from circles BUIX and BPUFY that

$$
\begin{aligned}
\angle(I U, U B)=\angle(I X, B X)=\angle\left(M^{\prime} X, B X\right)= & \frac{\pi-\angle A}{2} \\
& =\angle(E F, A F)=\angle(Y F, B F)=\angle(Y U, B U),
\end{aligned}
$$

so the points $Y, U$, and $I$ are collinear.
Let $E F$ meet $B C$ at $W$. We have

$$
\angle(I Y, Y W)=\angle(U Y, F Y)=\angle(U B, F B)=\angle(R F, A F)=\angle(C I, C W)
$$

so the points $W, Y, I$, and $C$ are concyclic.

Similarly, if $V$ and $Z$ are the second meeting points of $\Omega_{C}$ with $A R$ and $E F$, we get that the 4 -tuples $(C, V, I, X)$ and $(B, I, Z, W)$ are both concyclic.

Step 3. Let $Q^{\prime}=C Y \cap B Z$. We will show that $Q^{\prime}=Q$.
First of all, we have

$$
\begin{aligned}
& \angle\left(Q^{\prime} Y, Q^{\prime} B\right)=\angle(C Y, Z B)=\angle(C Y, Z Y)+\angle(Z Y, B Z) \\
& =\angle(C I, I W)+\angle(I W, I B)=\angle(C I, I B)=\frac{\pi-\angle A}{2}=\angle(F Y, F B),
\end{aligned}
$$

so $Q^{\prime} \in \Omega_{B}$. Similarly, $Q^{\prime} \in \Omega_{C}$. Thus $Q^{\prime} \in \Omega_{B} \cap \Omega_{C}=\{P, Q\}$ and it remains to prove that $Q^{\prime} \neq P$. If we had $Q^{\prime}=P$, we would have $\angle(P Y, P Z)=\angle\left(Q^{\prime} Y, Q^{\prime} Z\right)=\angle(I C, I B)$. This would imply

$$
\angle(P Y, Y F)+\angle(E Z, Z P)=\angle(P Y, P Z)=\angle(I C, I B)=\angle(P E, P F),
$$

so circles $\Omega_{B}$ and $\Omega_{C}$ would be tangent at $P$. That is excluded in the problem conditions, so $Q^{\prime}=Q$.


Step 4. Now we are ready to show that $P, Q$, and $S$ are collinear.
Notice that $A$ and $D$ are the poles of $E W$ and $D W$ with respect to $\omega$, so $W$ is the pole of $A D$. Hence, $W I \perp A D$. Since $C I \perp D E$, this yields $\angle(I C, W I)=\angle(D E, D A)$. On the other hand, $D A$ is a symmedian in $\triangle D E F$, so $\angle(D E, D A)=\angle(D N, D F)=\angle(D S, D F)$. Therefore,

$$
\begin{aligned}
\angle(P S, P F)=\angle(D S, D F)=\angle(D E, D A)= & \angle(I C, I W) \\
& =\angle(Y C, Y W)=\angle(Y Q, Y F)=\angle(P Q, P F),
\end{aligned}
$$

which yields the desired collinearity.

G8. Let $\mathcal{L}$ be the set of all lines in the plane and let $f$ be a function that assigns to each line $\ell \in \mathcal{L}$ a point $f(\ell)$ on $\ell$. Suppose that for any point $X$, and for any three lines $\ell_{1}, \ell_{2}, \ell_{3}$ passing through $X$, the points $f\left(\ell_{1}\right), f\left(\ell_{2}\right), f\left(\ell_{3}\right)$ and $X$ lie on a circle.

Prove that there is a unique point $P$ such that $f(\ell)=P$ for any line $\ell$ passing through $P$.
(Australia)
Common remarks. The condition on $f$ is equivalent to the following: There is some function $g$ that assigns to each point $X$ a circle $g(X)$ passing through $X$ such that for any line $\ell$ passing through $X$, the point $f(\ell)$ lies on $g(X)$. (The function $g$ may not be uniquely defined for all points, if some points $X$ have at most one value of $f(\ell)$ other than $X$; for such points, an arbitrary choice is made.)

If there were two points $P$ and $Q$ with the given property, $f(P Q)$ would have to be both $P$ and $Q$, so there is at most one such point, and it will suffice to show that such a point exists.

Solution 1. We provide a complete characterisation of the functions satisfying the given condition.

Write $\angle\left(\ell_{1}, \ell_{2}\right)$ for the directed angle modulo $180^{\circ}$ between the lines $\ell_{1}$ and $\ell_{2}$. Given a point $P$ and an angle $\alpha \in\left(0,180^{\circ}\right)$, for each line $\ell$, let $\ell^{\prime}$ be the line through $P$ satisfying $\angle\left(\ell^{\prime}, \ell\right)=\alpha$, and let $h_{P, \alpha}(\ell)$ be the intersection point of $\ell$ and $\ell^{\prime}$. We will prove that there is some pair $(P, \alpha)$ such that $f$ and $h_{P, \alpha}$ are the same function. Then $P$ is the unique point in the problem statement.

Given an angle $\alpha$ and a point $P$, let a line $\ell$ be called $(P, \alpha)$-good if $f(\ell)=h_{P, \alpha}(\ell)$. Let a point $X \neq P$ be called $(P, \alpha)$-good if the circle $g(X)$ passes through $P$ and some point $Y \neq P, X$ on $g(X)$ satisfies $\angle(P Y, Y X)=\alpha$. It follows from this definition that if $X$ is $(P, \alpha)$ good then every point $Y \neq P, X$ of $g(X)$ satisfies this angle condition, so $h_{P, \alpha}(X Y)=Y$ for every $Y \in g(X)$. Equivalently, $f(\ell) \in\left\{X, h_{P, \alpha}(\ell)\right\}$ for each line $\ell$ passing through $X$. This shows the following lemma.
Lemma 1. If $X$ is $(P, \alpha)$-good and $\ell$ is a line passing through $X$ then either $f(\ell)=X$ or $\ell$ is $(P, \alpha)$-good.
Lemma 2. If $X$ and $Y$ are different $(P, \alpha)$-good points, then line $X Y$ is $(P, \alpha)$-good.
Proof. If $X Y$ is not $(P, \alpha)$-good then by the previous Lemma, $f(X Y)=X$ and similarly $f(X Y)=Y$, but clearly this is impossible as $X \neq Y$.

Lemma 3. If $\ell_{1}$ and $\ell_{2}$ are different $(P, \alpha)$-good lines which intersect at $X \neq P$, then either $f\left(\ell_{1}\right)=X$ or $f\left(\ell_{2}\right)=X$ or $X$ is $(P, \alpha)$-good.
Proof. If $f\left(\ell_{1}\right), f\left(\ell_{2}\right) \neq X$, then $g(X)$ is the circumcircle of $X, f\left(\ell_{1}\right)$ and $f\left(\ell_{2}\right)$. Since $\ell_{1}$ and $\ell_{2}$ are $(P, \alpha)$-good lines, the angles

$$
\angle\left(P f\left(\ell_{1}\right), f\left(\ell_{1}\right) X\right)=\angle\left(P f\left(\ell_{2}\right), f\left(\ell_{2}\right) X\right)=\alpha,
$$

so $P$ lies on $g(X)$. Hence, $X$ is $(P, \alpha)$-good.
Lemma 4. If $\ell_{1}, \ell_{2}$ and $\ell_{3}$ are different $(P, \alpha)$-good lines which intersect at $X \neq P$, then $X$ is $(P, \alpha)$-good.
Proof. This follows from the previous Lemma since at most one of the three lines $\ell_{i}$ can satisfy $f\left(\ell_{i}\right)=X$ as the three lines are all $(P, \alpha)$-good.
Lemma 5. If $A B C$ is a triangle such that $A, B, C, f(A B), f(A C)$ and $f(B C)$ are all different points, then there is some point $P$ and some angle $\alpha$ such that $A, B$ and $C$ are $(P, \alpha)$-good points and $A B, B C$ and $C A$ are $(P, \alpha)$-good lines.


Proof. Let $D, E, F$ denote the points $f(B C), f(A C), f(A B)$, respectively. Then $g(A)$, $g(B)$ and $g(C)$ are the circumcircles of $A E F, B D F$ and $C D E$, respectively. Let $P \neq F$ be the second intersection of circles $g(A)$ and $g(B)$ (or, if these circles are tangent at $F$, then $P=F$ ). By Miquel's theorem (or an easy angle chase), $g(C)$ also passes through $P$. Then by the cyclic quadrilaterals, the directed angles

$$
\angle(P D, D C)=\angle(P F, F B)=\angle(P E, E A)=\alpha
$$

for some angle $\alpha$. Hence, lines $A B, B C$ and $C A$ are all $(P, \alpha)$-good, so by Lemma $3, A, B$ and $C$ are $(P, \alpha)$-good. (In the case where $P=D$, the line $P D$ in the equation above denotes the line which is tangent to $g(B)$ at $P=D$. Similar definitions are used for $P E$ and $P F$ in the cases where $P=E$ or $P=F$.)

Consider the set $\Omega$ of all points $(x, y)$ with integer coordinates $1 \leqslant x, y \leqslant 1000$, and consider the set $L_{\Omega}$ of all horizontal, vertical and diagonal lines passing through at least one point in $\Omega$. A simple counting argument shows that there are 5998 lines in $L_{\Omega}$. For each line $\ell$ in $L_{\Omega}$ we colour the point $f(\ell)$ red. Then there are at most 5998 red points. Now we partition the points in $\Omega$ into 10000 ten by ten squares. Since there are at most 5998 red points, at least one of these squares $\Omega_{10}$ contains no red points. Let $(m, n)$ be the bottom left point in $\Omega_{10}$. Then the triangle with vertices $(m, n),(m+1, n)$ and $(m, n+1)$ satisfies the condition of Lemma 5 , so these three vertices are all $(P, \alpha)$-good for some point $P$ and angle $\alpha$, as are the lines joining them. From this point on, we will simply call a point or line good if it is $(P, \alpha)$-good for this particular pair $(P, \alpha)$. Now by Lemma 1, the line $x=m+1$ is good, as is the line $y=n+1$. Then Lemma 3 implies that $(m+1, n+1)$ is good. By applying these two lemmas repeatedly, we can prove that the line $x+y=m+n+2$ is good, then the points $(m, n+2)$ and $(m+2, n)$ then the lines $x=m+2$ and $y=n+2$, then the points $(m+2, n+1),(m+1, n+2)$ and $(m+2, n+2)$ and so on until we have prove that all points in $\Omega_{10}$ are good.

Now we will use this to prove that every point $S \neq P$ is good. Since $g(S)$ is a circle, it passes through at most two points of $\Omega_{10}$ on any vertical line, so at most 20 points in total. Moreover, any line $\ell$ through $S$ intersects at most 10 points in $\Omega_{10}$. Hence, there are at least eight lines $\ell$ through $S$ which contain a point $Q$ in $\Omega_{10}$ which is not on $g(S)$. Since $Q$ is not on $g(S)$, the point $f(\ell) \neq Q$. Hence, by Lemma 1 , the line $\ell$ is good. Hence, at least eight good lines pass through $S$, so by Lemma 4, the point $S$ is good. Hence, every point $S \neq P$ is good, so by Lemma 2, every line is good. In particular, every line $\ell$ passing through $P$ is good, and therefore satisfies $f(\ell)=P$, as required.

Solution 2. Note that for any distinct points $X, Y$, the circles $g(X)$ and $g(Y)$ meet on $X Y$ at the point $f(X Y) \in g(X) \cap g(Y) \cap(X Y)$. We write $s(X, Y)$ for the second intersection point of circles $g(X)$ and $g(Y)$.
Lemma 1. Suppose that $X, Y$ and $Z$ are not collinear, and that $f(X Y) \notin\{X, Y\}$ and similarly for $Y Z$ and $Z X$. Then $s(X, Y)=s(Y, Z)=s(Z, X)$.
Proof. The circles $g(X), g(Y)$ and $g(Z)$ through the vertices of triangle $X Y Z$ meet pairwise on the corresponding edges (produced). By Miquel's theorem, the second points of intersection of any two of the circles coincide. (See the diagram for Lemma 5 of Solution 1.)

Now pick any line $\ell$ and any six different points $Y_{1}, \ldots, Y_{6}$ on $\ell \backslash\{f(\ell)\}$. Pick a point $X$ not on $\ell$ or any of the circles $g\left(Y_{i}\right)$. Reordering the indices if necessary, we may suppose that $Y_{1}, \ldots, Y_{4}$ do not lie on $g(X)$, so that $f\left(X Y_{i}\right) \notin\left\{X, Y_{i}\right\}$ for $1 \leqslant i \leqslant 4$. By applying the above lemma to triangles $X Y_{i} Y_{j}$ for $1 \leqslant i<j \leqslant 4$, we find that the points $s\left(Y_{i}, Y_{j}\right)$ and $s\left(X, Y_{i}\right)$ are all equal, to point $O$ say. Note that either $O$ does not lie on $\ell$, or $O=f(\ell)$, since $O \in g\left(Y_{i}\right)$.

Now consider an arbitrary point $X^{\prime}$ not on $\ell$ or any of the $\operatorname{circles~} g\left(Y_{i}\right)$ for $1 \leqslant i \leqslant 4$. As above, we see that there are two indices $1 \leqslant i<j \leqslant 4$ such that $Y_{i}$ and $Y_{j}$ do not lie on $g\left(X^{\prime}\right)$. By applying the above lemma to triangle $X^{\prime} Y_{i} Y_{j}$ we see that $s\left(X^{\prime}, Y_{i}\right)=O$, and in particular $g\left(X^{\prime}\right)$ passes through $O$.

We will now show that $f\left(\ell^{\prime}\right)=O$ for all lines $\ell^{\prime}$ through $O$. By the above note, we may assume that $\ell^{\prime} \neq \ell$. Consider a variable point $X^{\prime} \in \ell^{\prime} \backslash\{O\}$ not on $\ell$ or any of the circles $g\left(Y_{i}\right)$ for $1 \leqslant i \leqslant 4$. We know that $f\left(\ell^{\prime}\right) \in g\left(X^{\prime}\right) \cap \ell^{\prime}=\left\{X^{\prime}, O\right\}$. Since $X^{\prime}$ was suitably arbitrary, we have $f\left(\ell^{\prime}\right)=O$ as desired.

Solution 3. Notice that, for any two different points $X$ and $Y$, the point $f(X Y)$ lies on both $g(X)$ and $g(Y)$, so any two such circles meet in at least one point. We refer to two circles as cutting only in the case where they cross, and so meet at exactly two points, thus excluding the cases where they are tangent or are the same circle.
Lemma 1. Suppose there is a point $P$ such that all circles $g(X)$ pass through $P$. Then $P$ has the given property.
Proof. Consider some line $\ell$ passing through $P$, and suppose that $f(\ell) \neq P$. Consider some $X \in \ell$ with $X \neq P$ and $X \neq f(\ell)$. Then $g(X)$ passes through all of $P, f(\ell)$ and $X$, but those three points are collinear, a contradiction.

Lemma 2. Suppose that, for all $\epsilon>0$, there is a point $P_{\epsilon}$ with $g\left(P_{\epsilon}\right)$ of radius at most $\epsilon$. Then there is a point $P$ with the given property.
Proof. Consider a sequence $\epsilon_{i}=2^{-i}$ and corresponding points $P_{\epsilon_{i}}$. Because the two circles $g\left(P_{\epsilon_{i}}\right)$ and $g\left(P_{\epsilon_{j}}\right)$ meet, the distance between $P_{\epsilon_{i}}$ and $P_{\epsilon_{j}}$ is at most $2^{1-i}+2^{1-j}$. As $\sum_{i} \epsilon_{i}$ converges, these points converge to some point $P$. For all $\epsilon>0$, the point $P$ has distance at most $2 \epsilon$ from $P_{\epsilon}$, and all circles $g(X)$ pass through a point with distance at most $2 \epsilon$ from $P_{\epsilon}$, so distance at most $4 \epsilon$ from $P$. A circle that passes distance at most $4 \epsilon$ from $P$ for all $\epsilon>0$ must pass through $P$, so by Lemma 1 the point $P$ has the given property.

Lemma 3. Suppose no two of the circles $g(X)$ cut. Then there is a point $P$ with the given property.
Proof. Consider a circle $g(X)$ with centre $Y$. The circle $g(Y)$ must meet $g(X)$ without cutting it, so has half the radius of $g(X)$. Repeating this argument, there are circles with arbitrarily small radius and the result follows by Lemma 2.

Lemma 4. Suppose there are six different points $A, B_{1}, B_{2}, B_{3}, B_{4}, B_{5}$ such that no three are collinear, no four are concyclic, and all the circles $g\left(B_{i}\right)$ cut pairwise at $A$. Then there is a point $P$ with the given property.
Proof. Consider some line $\ell$ through $A$ that does not pass through any of the $B_{i}$ and is not tangent to any of the $g\left(B_{i}\right)$. Fix some direction along that line, and let $X_{\epsilon}$ be the point on $\ell$ that has distance $\epsilon$ from $A$ in that direction. In what follows we consider only those $\epsilon$ for which $X_{\epsilon}$ does not lie on any $g\left(B_{i}\right)$ (this restriction excludes only finitely many possible values of $\epsilon$ ).

Consider the circle $g\left(X_{\epsilon}\right)$. Because no four of the $B_{i}$ are concyclic, at most three of them lie on this circle, so at least two of them do not. There must be some sequence of $\epsilon \rightarrow 0$ such that it is the same two of the $B_{i}$ for all $\epsilon$ in that sequence, so now restrict attention to that sequence, and suppose without loss of generality that $B_{1}$ and $B_{2}$ do not lie on $g\left(X_{\epsilon}\right)$ for any $\epsilon$ in that sequence.

Then $f\left(X_{\epsilon} B_{1}\right)$ is not $B_{1}$, so must be the other point of intersection of $X_{\epsilon} B_{1}$ with $g\left(B_{1}\right)$, and the same applies with $B_{2}$. Now consider the three points $X_{\epsilon}, f\left(X_{\epsilon} B_{1}\right)$ and $f\left(X_{\epsilon} B_{2}\right)$. As $\epsilon \rightarrow 0$, the angle at $X_{\epsilon}$ tends to $\angle B_{1} A B_{2}$ or $180^{\circ}-\angle B_{1} A B_{2}$, which is not 0 or $180^{\circ}$ because no three of the points were collinear. All three distances between those points are bounded above by constant multiples of $\epsilon$ (in fact, if the triangle is scaled by a factor of $1 / \epsilon$, it tends to a fixed triangle). Thus the circumradius of those three points, which is the radius of $g\left(X_{\epsilon}\right)$, is also bounded above by a constant multiple of $\epsilon$, and so the result follows by Lemma 2 .

Lemma 5. Suppose there are two points $A$ and $B$ such that $g(A)$ and $g(B)$ cut. Then there is a point $P$ with the given property.

Proof. Suppose that $g(A)$ and $g(B)$ cut at $C$ and $D$. One of those points, without loss of generality $C$, must be $f(A B)$, and so lie on the line $A B$. We now consider two cases, according to whether $D$ also lies on that line.

Case 1: $D$ does not lie on that line.
In this case, consider a sequence of $X_{\epsilon}$ at distance $\epsilon$ from $D$, tending to $D$ along some line that is not a tangent to either circle, but perturbed slightly (by at most $\epsilon^{2}$ ) to ensure that no three of the points $A, B$ and $X_{\epsilon}$ are collinear and no four are concyclic.

Consider the points $f\left(X_{\epsilon} A\right)$ and $f\left(X_{\epsilon} B\right)$, and the circles $g\left(X_{\epsilon}\right)$ on which they lie. The point $f\left(X_{\epsilon} A\right)$ might be either $A$ or the other intersection of $X_{\epsilon} A$ with the circle $g(A)$, and the same applies for $B$. If, for some sequence of $\epsilon \rightarrow 0$, both those points are the other point of intersection, the same argument as in the proof of Lemma 4 applies to find arbitrarily small circles. Otherwise, we have either infinitely many of those circles passing through $A$, or infinitely many passing through $B$; without loss of generality, suppose infinitely many through $A$.

We now show we can find five points $B_{i}$ satisfying the conditions of Lemma 4 (together with $A$ ). Let $B_{1}$ be any of the $X_{\epsilon}$ for which $g\left(X_{\epsilon}\right)$ passes through $A$. Then repeat the following four times, for $2 \leqslant i \leqslant 5$.

Consider some line $\ell=X_{\epsilon} A$ (different from those considered for previous $i$ ) that is not tangent to any of the $g\left(B_{j}\right)$ for $j<i$, and is such that $f(\ell)=A$, so $g(Y)$ passes through $A$ for all $Y$ on that line. If there are arbitrarily small circles $g(Y)$ we are done by Lemma 2, so the radii of such circles must be bounded below. But as $Y \rightarrow A$, along any line not tangent to $g\left(B_{j}\right)$, the radius of a circle through $Y$ and tangent to $g\left(B_{j}\right)$ at $A$ tends to 0 . So there must be some $Y$ such that $g(Y)$ cuts $g\left(B_{j}\right)$ at $A$ rather than being tangent to it there, for all of the previous $B_{j}$, and we may also pick it such that no three of the $B_{i}$ and $A$ are collinear and no four are concyclic. Let $B_{i}$ be this $Y$. Now the result follows by Lemma 4 .

## Case 2: $D$ does lie on that line.

In this case, we follow a similar argument, but the sequence of $X_{\epsilon}$ needs to be slightly different. $C$ and $D$ both lie on the line $A B$, so one must be $A$ and the other must be $B$. Consider a sequence of $X_{\epsilon}$ tending to $B$. Rather than tending to $B$ along a straight line (with small perturbations), let the sequence be such that all the points are inside the two circles, with the angle between $X_{\epsilon} B$ and the tangent to $g(B)$ at $B$ tending to 0 .

Again consider the points $f\left(X_{\epsilon} A\right)$ and $f\left(X_{\epsilon} B\right)$. If, for some sequence of $\epsilon \rightarrow 0$, both those points are the other point of intersection with the respective circles, we see that the angle at $X_{\epsilon}$ tends to the angle between $A B$ and the tangent to $g(B)$ at $B$, which is not 0 or $180^{\circ}$, while the distances tend to 0 (although possibly slower than any multiple of $\epsilon$ ), so we have arbitrarily small circumradii and the result follows by Lemma 2. Otherwise, we have either infinitely many of the circles $g\left(X_{\epsilon}\right)$ passing through $A$, or infinitely many passing through $B$, and the same argument as in the previous case enables us to reduce to Lemma 4.

Lemmas 3 and 5 together cover all cases, and so the required result is proved.

Comment. From the property that all circles $g(X)$ pass through the same point $P$, it is possible to deduce that the function $f$ has the form given in Solution 1. For any line $\ell$ not passing through $P$ we may define a corresponding angle $\alpha(\ell)$, which we must show is the same for all such lines. For any point $X \neq P$, with at least one line $\ell$ through $X$ and not through $P$, such that $f(\ell) \neq X$, this angle must be equal for all such lines through $X$ (by (directed) angles in the same segment of $g(X)$ ).

Now consider all horizontal and all vertical lines not through $P$. For any pair consisting of a horizontal line $\ell_{1}$ and a vertical line $\ell_{2}$, we have $\alpha\left(\ell_{1}\right)=\alpha\left(\ell_{2}\right)$ unless $f\left(\ell_{1}\right)$ or $f\left(\ell_{2}\right)$ is the point of intersection of those lines. Consider the bipartite graph whose vertices are those lines and where an edge joins a horizontal and a vertical line with the same value of $\alpha$. Considering a subgraph induced by $n$ horizontal and $n$ vertical lines, it must have at least $n^{2}-2 n$ edges, so some horizontal line has edges to at least $n-2$ of the vertical lines. Thus, in the original graph, all but at most two of the vertical lines have the same value of $\alpha$, and likewise all but at most two of the horizontal lines have the same value of $\alpha$, and, restricting attention to suitable subsets of those lines, we see that this value must be the same for the vertical lines and for the horizontal lines.

But now we can extend this to all vertical and horizontal lines not through $P$ (and thus to lines in other directions as well, since the only requirement for 'vertical' and 'horizontal' above is that they are any two nonparallel directions). Consider any horizontal line $\ell_{1}$ not passing through $P$, and we wish to show that $\alpha\left(\ell_{1}\right)$ has the same value $\alpha$ it has for all but at most two lines not through $P$ in any direction. Indeed, we can deduce this by considering the intersection with any but at most five of the vertical lines: the only ones to exclude are the one passing through $P$, the one passing through $f\left(\ell_{1}\right)$, at most two such that $\alpha(\ell) \neq \alpha$, and the one passing through $h_{P, \alpha}\left(\ell_{1}\right)$ (defined as in Solution 1). So all lines $\ell$ not passing through $P$ have the same value of $\alpha(\ell)$.

Solution 4. For any point $X$, denote by $t(X)$ the line tangent to $g(X)$ at $X$; notice that $f(t(X))=X$, so $f$ is surjective.

Step 1: We find a point $P$ for which there are at least two different lines $p_{1}$ and $p_{2}$ such that $f\left(p_{i}\right)=P$.

Choose any point $X$. If $X$ does not have this property, take any $Y \in g(X) \backslash\{X\}$; then $f(X Y)=Y$. If $Y$ does not have the property, $t(Y)=X Y$, and the circles $g(X)$ and $g(Y)$ meet again at some point $Z$. Then $f(X Z)=Z=f(Y Z)$, so $Z$ has the required property.

We will show that $P$ is the desired point. From now on, we fix two different lines $p_{1}$ and $p_{2}$ with $f\left(p_{1}\right)=f\left(p_{2}\right)=P$. Assume for contradiction that $f(\ell)=Q \neq P$ for some line $\ell$ through $P$. We fix $\ell$, and note that $Q \in g(P)$.

Step 2: We prove that $P \in g(Q)$.
Take an arbitrary point $X \in \ell \backslash\{P, Q\}$. Two cases are possible for the position of $t(X)$ in relation to the $p_{i}$; we will show that each case (and subcase) occurs for only finitely many positions of $X$, yielding a contradiction.

Case 2.1: $t(X)$ is parallel to one of the $p_{i}$; say, to $p_{1}$.
Let $t(X)$ cross $p_{2}$ at $R$. Then $g(R)$ is the circle $(P R X)$, as $f(R P)=P$ and $f(R X)=X$. Let $R Q$ cross $g(R)$ again at $S$. Then $f(R Q) \in\{R, S\} \cap g(Q)$, so $g(Q)$ contains one of the points $R$ and $S$.

If $R \in g(Q)$, then $R$ is one of finitely many points in the intersection $g(Q) \cap p_{2}$, and each of them corresponds to a unique position of $X$, since $R X$ is parallel to $p_{1}$.

If $S \in g(Q)$, then $\angle(Q S, S P)=\angle(R S, S P)=\angle(R X, X P)=\angle\left(p_{1}, \ell\right)$, so $\angle(Q S, S P)$ is constant for all such points $X$, and all points $S$ obtained in such a way lie on one circle $\gamma$ passing through $P$ and $Q$. Since $g(Q)$ does not contain $P$, it is different from $\gamma$, so there are only finitely many points $S$. Each of them uniquely determines $R$ and thus $X$.


So, Case 2.1 can occur for only finitely many points $X$.
Case 2.2: $t(X)$ crosses $p_{1}$ and $p_{2}$ at $R_{1}$ and $R_{2}$, respectively.
Clearly, $R_{1} \neq R_{2}$, as $t(X)$ is the tangent to $g(X)$ at $X$, and $g(X)$ meets $\ell$ only at $X$ and $Q$. Notice that $g\left(R_{i}\right)$ is the circle $\left(P X R_{i}\right)$. Let $R_{i} Q$ meet $g\left(R_{i}\right)$ again at $S_{i}$; then $S_{i} \neq Q$, as $g\left(R_{i}\right)$ meets $\ell$ only at $P$ and $X$. Then $f\left(R_{i} Q\right) \in\left\{R_{i}, S_{i}\right\}$, and we distinguish several subcases.


Subcase 2.2.1: $f\left(R_{1} Q\right)=S_{1}, f\left(R_{2} Q\right)=S_{2}$; so $S_{1}, S_{2} \in g(Q)$.
In this case we have $0=\angle\left(R_{1} X, X P\right)+\angle\left(X P, R_{2} X\right)=\angle\left(R_{1} S_{1}, S_{1} P\right)+\angle\left(S_{2} P, S_{2} R_{2}\right)=$ $\angle\left(Q S_{1}, S_{1} P\right)+\angle\left(S_{2} P, S_{2} Q\right)$, which shows $P \in g(Q)$.

Subcase 2.2.2: $f\left(R_{1} Q\right)=R_{1}, f\left(R_{2} Q\right)=R_{2}$; so $R_{1}, R_{2} \in g(Q)$.
This can happen for at most four positions of $X$ - namely, at the intersections of $\ell$ with a line of the form $K_{1} K_{2}$, where $K_{i} \in g(Q) \cap p_{i}$.

Subcase 2.2.3: $f\left(R_{1} Q\right)=S_{1}, f\left(R_{2} Q\right)=R_{2}$ (the case $f\left(R_{1} Q\right)=R_{1}, f\left(R_{2} Q\right)=S_{2}$ is similar).
In this case, there are at most two possible positions for $R_{2}$ - namely, the meeting points of $g(Q)$ with $p_{2}$. Consider one of them. Let $X$ vary on $\ell$. Then $R_{1}$ is the projection of $X$ to $p_{1}$ via $R_{2}, S_{1}$ is the projection of $R_{1}$ to $g(Q)$ via $Q$. Finally, $\angle\left(Q S_{1}, S_{1} X\right)=\angle\left(R_{1} S_{1}, S_{1} X\right)=$ $\angle\left(R_{1} P, P X\right)=\angle\left(p_{1}, \ell\right) \neq 0$, so $X$ is obtained by a fixed projective transform $g(Q) \rightarrow \ell$ from $S_{1}$. So, if there were three points $X$ satisfying the conditions of this subcase, the composition of the three projective transforms would be the identity. But, if we apply it to $X=Q$, we successively get some point $R_{1}^{\prime}$, then $R_{2}$, and then some point different from $Q$, a contradiction.

Thus Case 2.2 also occurs for only finitely many points $X$, as desired.
Step 3: We show that $f(P Q)=P$, as desired.
The argument is similar to that in Step 2, with the roles of $Q$ and $X$ swapped. Again, we show that there are only finitely many possible positions for a point $X \in \ell \backslash\{P, Q\}$, which is absurd.
Case 3.1: $t(Q)$ is parallel to one of the $p_{i}$; say, to $p_{1}$.
Let $t(Q)$ cross $p_{2}$ at $R$; then $g(R)$ is the circle (PRQ). Let $R X$ cross $g(R)$ again at $S$. Then $f(R X) \in\{R, S\} \cap g(X)$, so $g(X)$ contains one of the points $R$ and $S$.


Subcase 3.1.1: $S=f(R X) \in g(X)$.
We have $\angle(t(X), Q X)=\angle(S X, S Q)=\angle(S R, S Q)=\angle(P R, P Q)=\angle\left(p_{2}, \ell\right)$. Hence $t(X) \| p_{2}$. Now we recall Case 2.1: we let $t(X)$ cross $p_{1}$ at $R^{\prime}$, so $g\left(R^{\prime}\right)=\left(P R^{\prime} X\right)$, and let $R^{\prime} Q$ meet $g\left(R^{\prime}\right)$ again at $S^{\prime}$; notice that $S^{\prime} \neq Q$. Excluding one position of $X$, we may assume that $R^{\prime} \notin g(Q)$, so $R^{\prime} \neq f\left(R^{\prime} Q\right)$. Therefore, $S^{\prime}=f\left(R^{\prime} Q\right) \in g(Q)$. But then, as in Case 2.1, we get $\angle(t(Q), P Q)=\angle\left(Q S^{\prime}, S^{\prime} P\right)=\angle\left(R^{\prime} X, X P\right)=\angle\left(p_{2}, \ell\right)$. This means that $t(Q)$ is parallel to $p_{2}$, which is impossible.
Subcase 3.1.2: $R=f(R X) \in g(X)$.
In this case, we have $\angle(t(X), \ell)=\angle(R X, R Q)=\angle\left(R X, p_{1}\right)$. Again, let $R^{\prime}=t(X) \cap p_{1}$; this point exists for all but at most one position of $X$. Then $g\left(R^{\prime}\right)=\left(R^{\prime} X P\right)$; let $R^{\prime} Q$ meet $g\left(R^{\prime}\right)$ again at $S^{\prime}$. Due to $\angle\left(R^{\prime} X, X R\right)=\angle(Q X, Q R)=\angle\left(\ell, p_{1}\right), R^{\prime}$ determines $X$ in at most two ways, so for all but finitely many positions of $X$ we have $R^{\prime} \notin g(Q)$. Therefore, for those positions we have $S^{\prime}=f\left(R^{\prime} Q\right) \in g(Q)$. But then $\angle\left(R X, p_{1}\right)=\angle\left(R^{\prime} X, X P\right)=\angle\left(R^{\prime} S^{\prime}, S^{\prime} P\right)=$ $\angle\left(Q S^{\prime}, S^{\prime} P\right)=\angle(t(Q), Q P)$ is fixed, so this case can hold only for one specific position of $X$ as well.

Thus, in Case 3.1, there are only finitely many possible positions of $X$, yielding a contradiction.

Case 3.2: $t(Q)$ crosses $p_{1}$ and $p_{2}$ at $R_{1}$ and $R_{2}$, respectively.
By Step $2, R_{1} \neq R_{2}$. Notice that $g\left(R_{i}\right)$ is the circle $\left(P Q R_{i}\right)$. Let $R_{i} X$ meet $g\left(R_{i}\right)$ at $S_{i}$; then $S_{i} \neq X$. Then $f\left(R_{i} X\right) \in\left\{R_{i}, S_{i}\right\}$, and we distinguish several subcases.


Subcase 3.2.1: $f\left(R_{1} X\right)=S_{1}$ and $f\left(R_{2} X\right)=S_{2}$, so $S_{1}, S_{2} \in g(X)$.
As in Subcase 2.2.1, we have $0=\angle\left(R_{1} Q, Q P\right)+\angle\left(Q P, R_{2} Q\right)=\angle\left(X S_{1}, S_{1} P\right)+\angle\left(S_{2} P, S_{2} X\right)$, which shows $P \in g(X)$. But $X, Q \in g(X)$ as well, so $g(X)$ meets $\ell$ at three distinct points, which is absurd.

Subcase 3.2.2: $f\left(R_{1} X\right)=R_{1}, f\left(R_{2} X\right)=R_{2}$, so $R_{1}, R_{2} \in g(X)$.
Now three distinct collinear points $R_{1}, R_{2}$, and $Q$ belong to $g(X)$, which is impossible.
Subcase 3.2.3: $f\left(R_{1} X\right)=S_{1}, f\left(R_{2} X\right)=R_{2}$ (the case $f\left(R_{1} X\right)=R_{1}, f\left(R_{2} X\right)=S_{2}$ is similar).
We have $\angle\left(X R_{2}, R_{2} Q\right)=\angle\left(X S_{1}, S_{1} Q\right)=\angle\left(R_{1} S_{1}, S_{1} Q\right)=\angle\left(R_{1} P, P Q\right)=\angle\left(p_{1}, \ell\right)$, so this case can occur for a unique position of $X$.

Thus, in Case 3.2, there is only a unique position of $X$, again yielding the required contradiction.

## Number Theory

N1. Find all pairs ( $m, n$ ) of positive integers satisfying the equation

$$
\begin{equation*}
\left(2^{n}-1\right)\left(2^{n}-2\right)\left(2^{n}-4\right) \cdots\left(2^{n}-2^{n-1}\right)=m! \tag{1}
\end{equation*}
$$

(El Salvador)
Answer: The only such pairs are $(1,1)$ and $(3,2)$.
Common remarks. In all solutions, for any prime $p$ and positive integer $N$, we will denote by $v_{p}(N)$ the exponent of the largest power of $p$ that divides $N$. The left-hand side of (1) will be denoted by $L_{n}$; that is, $L_{n}=\left(2^{n}-1\right)\left(2^{n}-2\right)\left(2^{n}-4\right) \cdots\left(2^{n}-2^{n-1}\right)$.

Solution 1. We will get an upper bound on $n$ from the speed at which $v_{2}\left(L_{n}\right)$ grows.
From

$$
L_{n}=\left(2^{n}-1\right)\left(2^{n}-2\right) \cdots\left(2^{n}-2^{n-1}\right)=2^{1+2+\cdots+(n-1)}\left(2^{n}-1\right)\left(2^{n-1}-1\right) \cdots\left(2^{1}-1\right)
$$

we read

$$
v_{2}\left(L_{n}\right)=1+2+\cdots+(n-1)=\frac{n(n-1)}{2} .
$$

On the other hand, $v_{2}(m!)$ is expressed by the Legendre formula as

$$
v_{2}(m!)=\sum_{i=1}^{\infty}\left\lfloor\frac{m}{2^{i}}\right\rfloor .
$$

As usual, by omitting the floor functions,

$$
v_{2}(m!)<\sum_{i=1}^{\infty} \frac{m}{2^{i}}=m .
$$

Thus, $L_{n}=m$ ! implies the inequality

$$
\begin{equation*}
\frac{n(n-1)}{2}<m . \tag{2}
\end{equation*}
$$

In order to obtain an opposite estimate, observe that

$$
L_{n}=\left(2^{n}-1\right)\left(2^{n}-2\right) \cdots\left(2^{n}-2^{n-1}\right)<\left(2^{n}\right)^{n}=2^{n^{2}} .
$$

We claim that

$$
\begin{equation*}
2^{n^{2}}<\left(\frac{n(n-1)}{2}\right)!\text { for } n \geqslant 6 \tag{3}
\end{equation*}
$$

For $n=6$ the estimate (3) is true because $2^{6^{2}}<6.9 \cdot 10^{10}$ and $\left(\frac{n(n-1)}{2}\right)$ ! $=15!>1.3 \cdot 10^{12}$.
For $n \geqslant 7$ we prove (3) by the following inequalities:

$$
\begin{aligned}
\left(\frac{n(n-1)}{2}\right)! & =15!\cdot 16 \cdot 17 \cdots \frac{n(n-1)}{2}>2^{36} \cdot 16^{\frac{n(n-1)}{2}-15} \\
& =2^{2 n(n-1)-24}=2^{n^{2}} \cdot 2^{n(n-2)-24}>2^{n^{2}} .
\end{aligned}
$$

Putting together (2) and (3), for $n \geqslant 6$ we get a contradiction, since

$$
L_{n}<2^{n^{2}}<\left(\frac{n(n-1)}{2}\right)!<m!=L_{n}
$$

Hence $n \geqslant 6$ is not possible.
Checking manually the cases $n \leqslant 5$ we find

$$
\begin{gathered}
L_{1}=1=1!, \quad L_{2}=6=3!, \quad 5!<L_{3}=168<6!, \\
7!<L_{4}=20160<8!\quad \text { and } \quad 10!<L_{5}=9999360<11!
\end{gathered}
$$

So, there are two solutions:

$$
(m, n) \in\{(1,1),(3,2)\} .
$$

Solution 2. Like in the previous solution, the cases $n=1,2,3,4$ are checked manually. We will exclude $n \geqslant 5$ by considering the exponents of 3 and 31 in (1).

For odd primes $p$ and distinct integers $a, b$, coprime to $p$, with $p \mid a-b$, the Lifting The Exponent lemma asserts that

$$
v_{p}\left(a^{k}-b^{k}\right)=v_{p}(a-b)+v_{p}(k) .
$$

Notice that 3 divides $2^{k}-1$ if only if $k$ is even; moreover, by the Lifting The Exponent lemma we have

$$
v_{3}\left(2^{2 k}-1\right)=v_{3}\left(4^{k}-1\right)=1+v_{3}(k)=v_{3}(3 k) .
$$

Hence,

$$
v_{3}\left(L_{n}\right)=\sum_{2 k \leqslant n} v_{3}\left(4^{k}-1\right)=\sum_{k \leqslant\left\lfloor\frac{n}{2}\right\rfloor} v_{3}(3 k) .
$$

Notice that the last expression is precisely the exponent of 3 in the prime factorisation of $\left(3\left\lfloor\frac{n}{2}\right\rfloor\right)$ !. Therefore

$$
\begin{gather*}
v_{3}(m!)=v_{3}\left(L_{n}\right)=v_{3}\left(\left(3\left\lfloor\frac{n}{2}\right\rfloor\right)!\right) \\
3\left\lfloor\frac{n}{2}\right\rfloor \leqslant m \leqslant 3\left\lfloor\frac{n}{2}\right\rfloor+2 . \tag{4}
\end{gather*}
$$

Suppose that $n \geqslant 5$. Note that every fifth factor in $L_{n}$ is divisible by $31=2^{5}-1$, and hence we have $v_{31}\left(L_{n}\right) \geqslant\left\lfloor\frac{n}{5}\right\rfloor$. Then

$$
\begin{equation*}
\frac{n}{10} \leqslant\left\lfloor\frac{n}{5}\right\rfloor \leqslant v_{31}\left(L_{n}\right)=v_{31}(m!)=\sum_{k=1}^{\infty}\left\lfloor\frac{m}{31^{k}}\right\rfloor<\sum_{k=1}^{\infty} \frac{m}{31^{k}}=\frac{m}{30} \tag{5}
\end{equation*}
$$

By combining (4) and (5),

$$
3 n<m \leqslant \frac{3 n}{2}+2
$$

so $n<\frac{4}{3}$ which is inconsistent with the inequality $n \geqslant 5$.
Comment 1. There are many combinations of the ideas above; for example combining (2) and (4) also provides $n<5$. Obviously, considering the exponents of any two primes in (1), or considering one prime and the magnitude orders lead to an upper bound on $n$ and $m$.

Comment 2. This problem has a connection to group theory. Indeed, the left-hand side is the order of the group $G L_{n}\left(\mathbb{F}_{2}\right)$ of invertible $n$-by- $n$ matrices with entries modulo 2 , while the right-hand side is the order of the symmetric group $S_{m}$ on $m$ elements. The result thus shows that the only possible isomorphisms between these groups are $G L_{1}\left(\mathbb{F}_{2}\right) \cong S_{1}$ and $G L_{2}\left(\mathbb{F}_{2}\right) \cong S_{3}$, and there are in fact isomorphisms in both cases. In general, $G L_{n}\left(\mathbb{F}_{2}\right)$ is a simple group for $n \geqslant 3$, as it is isomorphic to $P S L_{n}\left(\mathbb{F}_{2}\right)$.

There is also a near-solution of interest: the left-hand side for $n=4$ is half of the right-hand side when $m=8$; this turns out to correspond to an isomorphism $G L_{4}\left(\mathbb{F}_{2}\right) \cong A_{8}$ with the alternating group on eight elements.

However, while this indicates that the problem is a useful one, knowing group theory is of no use in solving it!

N2. Find all triples $(a, b, c)$ of positive integers such that $a^{3}+b^{3}+c^{3}=(a b c)^{2}$.
(Nigeria)
Answer: The solutions are (1, 2, 3) and its permutations.
Common remarks. Note that the equation is symmetric. In all solutions, we will assume without loss of generality that $a \geqslant b \geqslant c$, and prove that the only solution is $(a, b, c)=(3,2,1)$.

The first two solutions all start by proving that $c=1$.
Solution 1. We will start by proving that $c=1$. Note that

$$
3 a^{3} \geqslant a^{3}+b^{3}+c^{3}>a^{3} .
$$

So $3 a^{3} \geqslant(a b c)^{2}>a^{3}$ and hence $3 a \geqslant b^{2} c^{2}>a$. Now $b^{3}+c^{3}=a^{2}\left(b^{2} c^{2}-a\right) \geqslant a^{2}$, and so

$$
18 b^{3} \geqslant 9\left(b^{3}+c^{3}\right) \geqslant 9 a^{2} \geqslant b^{4} c^{4} \geqslant b^{3} c^{5},
$$

so $18 \geqslant c^{5}$ which yields $c=1$.
Now, note that we must have $a>b$, as otherwise we would have $2 b^{3}+1=b^{4}$ which has no positive integer solutions. So

$$
a^{3}-b^{3} \geqslant(b+1)^{3}-b^{3}>1
$$

and

$$
2 a^{3}>1+a^{3}+b^{3}>a^{3}
$$

which implies $2 a^{3}>a^{2} b^{2}>a^{3}$ and so $2 a>b^{2}>a$. Therefore

$$
4\left(1+b^{3}\right)=4 a^{2}\left(b^{2}-a\right) \geqslant 4 a^{2}>b^{4}
$$

so $4>b^{3}(b-4)$; that is, $b \leqslant 4$.
Now, for each possible value of $b$ with $2 \leqslant b \leqslant 4$ we obtain a cubic equation for $a$ with constant coefficients. These are as follows:

$$
\begin{array}{ll}
b=2: & \\
b=3: & a^{3}-4 a^{2}+9=0 \\
b=4: & \\
a^{3}-9 a^{2}+28=0 \\
a^{3}-16 a^{2}+65=0 .
\end{array}
$$

The only case with an integer solution for $a$ with $b \leqslant a$ is $b=2$, leading to $(a, b, c)=(3,2,1)$.
Comment 1.1. Instead of writing down each cubic equation explicitly, we could have just observed that $a^{2} \mid b^{3}+1$, and for each choice of $b$ checked each square factor of $b^{3}+1$ for $a^{2}$.

We could also have observed that, with $c=1$, the relation $18 b^{3} \geqslant b^{4} c^{4}$ becomes $b \leqslant 18$, and we can simply check all possibilities for $b$ (instead of working to prove that $b \leqslant 4$ ). This check becomes easier after using the factorisation $b^{3}+1=(b+1)\left(b^{2}-b+1\right)$ and observing that no prime besides 3 can divide both of the factors.

Comment 1.2. Another approach to finish the problem after establishing that $c \leqslant 1$ is to set $k=b^{2} c^{2}-a$, which is clearly an integer and must be positive as it is equal to $\left(b^{3}+c^{3}\right) / a^{2}$. Then we divide into cases based on whether $k=1$ or $k \geqslant 2$; in the first case, we have $b^{3}+1=a^{2}=\left(b^{2}-1\right)^{2}$ whose only positive root is $b=2$, and in the second case we have $b^{2} \leqslant 3 a$, and so

$$
b^{4} \leqslant(3 a)^{2} \leqslant \frac{9}{2}\left(k a^{2}\right)=\frac{9}{2}\left(b^{3}+1\right)
$$

which implies that $b \leqslant 4$.

Solution 2. Again, we will start by proving that $c=1$. Suppose otherwise that $c \geqslant 2$. We have $a^{3}+b^{3}+c^{3} \leqslant 3 a^{3}$, so $b^{2} c^{2} \leqslant 3 a$. Since $c \geqslant 2$, this tells us that $b \leqslant \sqrt{3 a / 4}$. As the right-hand side of the original equation is a multiple of $a^{2}$, we have $a^{2} \leqslant 2 b^{3} \leqslant 2(3 a / 4)^{3 / 2}$. In other words, $a \leqslant \frac{27}{16}<2$, which contradicts the assertion that $a \geqslant c \geqslant 2$. So there are no solutions in this case, and so we must have $c=1$.

Now, the original equation becomes $a^{3}+b^{3}+1=a^{2} b^{2}$. Observe that $a \geqslant 2$, since otherwise $a=b=1$ as $a \geqslant b$.

The right-hand side is a multiple of $a^{2}$, so the left-hand side must be as well. Thus, $b^{3}+1 \geqslant$ $a^{2}$. Since $a \geqslant b$, we also have

$$
b^{2}=a+\frac{b^{3}+1}{a^{2}} \leqslant 2 a+\frac{1}{a^{2}}
$$

and so $b^{2} \leqslant 2 a$ since $b^{2}$ is an integer. Thus $(2 a)^{3 / 2}+1 \geqslant b^{3}+1 \geqslant a^{2}$, from which we deduce $a \leqslant 8$.

Now, for each possible value of $a$ with $2 \leqslant a \leqslant 8$ we obtain a cubic equation for $b$ with constant coefficients. These are as follows:

$$
\begin{array}{ll}
a=2: & b^{3}-4 b^{2}+9=0 \\
a=3: & b^{3}-9 b^{2}+28=0 \\
a=4: & b^{3}-16 b^{2}+65=0 \\
a=5: & b^{3}-25 b^{2}+126=0 \\
a=6: & b^{3}-36 b^{2}+217=0 \\
a=7: & b^{3}-49 b^{2}+344=0 \\
a=8: & b^{3}-64 b^{2}+513=0 .
\end{array}
$$

The only case with an integer solution for $b$ with $a \geqslant b$ is $a=3$, leading to $(a, b, c)=(3,2,1)$.
Comment 2.1. As in Solution 1, instead of writing down each cubic equation explicitly, we could have just observed that $b^{2} \mid a^{3}+1$, and for each choice of $a$ checked each square factor of $a^{3}+1$ for $b^{2}$.

Comment 2.2. This solution does not require initially proving that $c=1$, in which case the bound would become $a \leqslant 108$. The resulting cases could, in principle, be checked by a particularly industrious student.

Solution 3. Set $k=\left(b^{3}+c^{3}\right) / a^{2} \leqslant 2 a$, and rewrite the original equation as $a+k=(b c)^{2}$. Since $b^{3}$ and $c^{3}$ are positive integers, we have $(b c)^{3} \geqslant b^{3}+c^{3}-1=k a^{2}-1$, so

$$
a+k \geqslant\left(k a^{2}-1\right)^{2 / 3}
$$

As in Comment 1.2, $k$ is a positive integer; for each value of $k \geqslant 1$, this gives us a polynomial inequality satisfied by $a$ :

$$
k^{2} a^{4}-a^{3}-5 k a^{2}-3 k^{2} a-\left(k^{3}-1\right) \leqslant 0 .
$$

We now prove that $a \leqslant 3$. Indeed,

$$
0 \geqslant \frac{k^{2} a^{4}-a^{3}-5 k a^{2}-3 k^{2} a-\left(k^{3}-1\right)}{k^{2}} \geqslant a^{4}-a^{3}-5 a^{2}-3 a-k \geqslant a^{4}-a^{3}-5 a^{2}-5 a
$$

which fails when $a \geqslant 4$.
This leaves ten triples with $3 \geqslant a \geqslant b \geqslant c \geqslant 1$, which may be checked manually to give $(a, b, c)=(3,2,1)$.

Solution 4. Again, observe that $b^{3}+c^{3}=a^{2}\left(b^{2} c^{2}-a\right)$, so $b \leqslant a \leqslant b^{2} c^{2}-1$.
We consider the function $f(x)=x^{2}\left(b^{2} c^{2}-x\right)$. It can be seen that that on the interval $\left[0, b^{2} c^{2}-1\right]$ the function $f$ is increasing if $x<\frac{2}{3} b^{2} c^{2}$ and decreasing if $x>\frac{2}{3} b^{2} c^{2}$. Consequently, it must be the case that

$$
b^{3}+c^{3}=f(a) \geqslant \min \left(f(b), f\left(b^{2} c^{2}-1\right)\right)
$$

First, suppose that $b^{3}+c^{3} \geqslant f\left(b^{2} c^{2}-1\right)$. This may be written $b^{3}+c^{3} \geqslant\left(b^{2} c^{2}-1\right)^{2}$, and so

$$
2 b^{3} \geqslant b^{3}+c^{3} \geqslant\left(b^{2} c^{2}-1\right)^{2}>b^{4} c^{4}-2 b^{2} c^{2} \geqslant b^{4} c^{4}-2 b^{3} c^{4} .
$$

Thus, $(b-2) c^{4}<2$, and the only solutions to this inequality have $(b, c)=(2,2)$ or $b \leqslant 3$ and $c=1$. It is easy to verify that the only case giving a solution for $a \geqslant b$ is $(a, b, c)=(3,2,1)$.

Otherwise, suppose that $b^{3}+c^{3}=f(a) \geqslant f(b)$. Then, we have

$$
2 b^{3} \geqslant b^{3}+c^{3}=a^{2}\left(b^{2} c^{2}-a\right) \geqslant b^{2}\left(b^{2} c^{2}-b\right) .
$$

Consequently $b c^{2} \leqslant 3$, with strict inequality in the case that $b \neq c$. Hence $c=1$ and $b \leqslant 2$. Both of these cases have been considered already, so we are done.

Comment 4.1. Instead of considering which of $f(b)$ and $f\left(b^{2} c^{2}-1\right)$ is less than $f(a)$, we may also proceed by explicitly dividing into cases based on whether $a \geqslant \frac{2}{3} b^{2} c^{2}$ or $a<\frac{2}{3} b^{2} c^{2}$. The first case may now be dealt with as follows. We have $b^{3} c^{3}+1 \geqslant b^{3}+c^{3}$ as $b^{3}$ and $c^{3}$ are positive integers, so we have

$$
b^{3} c^{3}+1 \geqslant b^{3}+c^{3} \geqslant a^{2} \geqslant \frac{4}{9} b^{4} c^{4} .
$$

This implies $b c \leqslant 2$, and hence $c=1$ and $b \leqslant 2$.

N3. We say that a set $S$ of integers is rootiful if, for any positive integer $n$ and any $a_{0}, a_{1}, \ldots, a_{n} \in S$, all integer roots of the polynomial $a_{0}+a_{1} x+\cdots+a_{n} x^{n}$ are also in $S$. Find all rootiful sets of integers that contain all numbers of the form $2^{a}-2^{b}$ for positive integers $a$ and $b$.
(Czech Republic)
Answer: The set $\mathbb{Z}$ of all integers is the only such rootiful set.
Solution 1. The set $\mathbb{Z}$ of all integers is clearly rootiful. We shall prove that any rootiful set $S$ containing all the numbers of the form $2^{a}-2^{b}$ for $a, b \in \mathbb{Z}_{>0}$ must be all of $\mathbb{Z}$.

First, note that $0=2^{1}-2^{1} \in S$ and $2=2^{2}-2^{1} \in S$. Now, $-1 \in S$, since it is a root of $2 x+2$, and $1 \in S$, since it is a root of $2 x^{2}-x-1$. Also, if $n \in S$ then $-n$ is a root of $x+n$, so it suffices to prove that all positive integers must be in $S$.

Now, we claim that any positive integer $n$ has a multiple in $S$. Indeed, suppose that $n=2^{\alpha} \cdot t$ for $\alpha \in \mathbb{Z}_{\geqslant 0}$ and $t$ odd. Then $t \mid 2^{\phi(t)}-1$, so $n \mid 2^{\alpha+\phi(t)+1}-2^{\alpha+1}$. Moreover, $2^{\alpha+\phi(t)+1}-2^{\alpha+1} \in S$, and so $S$ contains a multiple of every positive integer $n$.

We will now prove by induction that all positive integers are in $S$. Suppose that $0,1, \ldots, n-$ $1 \in S$; furthermore, let $N$ be a multiple of $n$ in $S$. Consider the base- $n$ expansion of $N$, say $N=a_{k} n^{k}+a_{k-1} n^{k-1}+\cdots+a_{1} n+a_{0}$. Since $0 \leqslant a_{i}<n$ for each $a_{i}$, we have that all the $a_{i}$ are in $S$. Furthermore, $a_{0}=0$ since $N$ is a multiple of $n$. Therefore, $a_{k} n^{k}+a_{k-1} n^{k-1}+\cdots+a_{1} n-N=0$, so $n$ is a root of a polynomial with coefficients in $S$. This tells us that $n \in S$, completing the induction.

Solution 2. As in the previous solution, we can prove that 0,1 and -1 must all be in any rootiful set $S$ containing all numbers of the form $2^{a}-2^{b}$ for $a, b \in \mathbb{Z}_{>0}$.

We show that, in fact, every integer $k$ with $|k|>2$ can be expressed as a root of a polynomial whose coefficients are of the form $2^{a}-2^{b}$. Observe that it suffices to consider the case where $k$ is positive, as if $k$ is a root of $a_{n} x^{n}+\cdots+a_{1} x+a_{0}=0$, then $-k$ is a root of $(-1)^{n} a_{n} x^{n}+\cdots-$ $a_{1} x+a_{0}=0$.

Note that

$$
\left(2^{a_{n}}-2^{b_{n}}\right) k^{n}+\cdots+\left(2^{a_{0}}-2^{b_{0}}\right)=0
$$

is equivalent to

$$
2^{a_{n}} k^{n}+\cdots+2^{a_{0}}=2^{b_{n}} k^{n}+\cdots+2^{b_{0}} .
$$

Hence our aim is to show that two numbers of the form $2^{a_{n}} k^{n}+\cdots+2^{a_{0}}$ are equal, for a fixed value of $n$. We consider such polynomials where every term $2^{a_{i}} k^{i}$ is at most $2 k^{n}$; in other words, where $2 \leqslant 2^{a_{i}} \leqslant 2 k^{n-i}$, or, equivalently, $1 \leqslant a_{i} \leqslant 1+(n-i) \log _{2} k$. Therefore, there must be $1+\left\lfloor(n-i) \log _{2} k\right\rfloor$ possible choices for $a_{i}$ satisfying these constraints.

The number of possible polynomials is then

$$
\prod_{i=0}^{n}\left(1+\left\lfloor(n-i) \log _{2} k\right\rfloor\right) \geqslant \prod_{i=0}^{n-1}(n-i) \log _{2} k=n!\left(\log _{2} k\right)^{n}
$$

where the inequality holds as $1+\lfloor x\rfloor \geqslant x$.
As there are $(n+1)$ such terms in the polynomial, each at most $2 k^{n}$, such a polynomial must have value at most $2 k^{n}(n+1)$. However, for large $n$, we have $n!\left(\log _{2} k\right)^{n}>2 k^{n}(n+1)$. Therefore there are more polynomials than possible values, so some two must be equal, as required.

N4. Let $\mathbb{Z}_{>0}$ be the set of positive integers. A positive integer constant $C$ is given. Find all functions $f: \mathbb{Z}_{>0} \rightarrow \mathbb{Z}_{>0}$ such that, for all positive integers $a$ and $b$ satisfying $a+b>C$,

$$
\begin{equation*}
a+f(b) \mid a^{2}+b f(a) \tag{*}
\end{equation*}
$$

(Croatia)
Answer: The functions satisfying (*) are exactly the functions $f(a)=k a$ for some constant $k \in \mathbb{Z}_{>0}$ (irrespective of the value of $C$ ).

Common remarks. It is easy to verify that the functions $f(a)=k a$ satisfy (*). Thus, in the proofs below, we will only focus on the converse implication: that condition $(*)$ implies that $f=k a$.

A common minor part of these solutions is the derivation of some relatively easy bounds on the function $f$. An upper bound is easily obtained by setting $a=1$ in (*), giving the inequality

$$
f(b) \leqslant b \cdot f(1)
$$

for all sufficiently large $b$. The corresponding lower bound is only marginally more difficult to obtain: substituting $b=1$ in the original equation shows that

$$
a+f(1) \mid\left(a^{2}+f(a)\right)-(a-f(1)) \cdot(a+f(1))=f(1)^{2}+f(a)
$$

for all sufficiently large $a$. It follows from this that one has the lower bound

$$
f(a) \geqslant a+f(1) \cdot(1-f(1)),
$$

again for all sufficiently large $a$.
Each of the following proofs makes use of at least one of these bounds.
Solution 1. First, we show that $b \mid f(b)^{2}$ for all $b$. To do this, we choose a large positive integer $n$ so that $n b-f(b) \geqslant C$. Setting $a=n b-f(b)$ in (*) then shows that

$$
n b \mid(n b-f(b))^{2}+b f(n b-f(b))
$$

so that $b \mid f(b)^{2}$ as claimed.
Now in particular we have that $p \mid f(p)$ for every prime $p$. If we write $f(p)=k(p) \cdot p$, then the bound $f(p) \leqslant f(1) \cdot p$ (valid for $p$ sufficiently large) shows that some value $k$ of $k(p)$ must be attained for infinitely many $p$. We will show that $f(a)=k a$ for all positive integers $a$. To do this, we substitute $b=p$ in $(*)$, where $p$ is any sufficiently large prime for which $k(p)=k$, obtaining

$$
a+k p \mid\left(a^{2}+p f(a)\right)-a(a+k p)=p f(a)-p k a .
$$

For suitably large $p$ we have $\operatorname{gcd}(a+k p, p)=1$, and hence we have

$$
a+k p \mid f(a)-k a .
$$

But the only way this can hold for arbitrarily large $p$ is if $f(a)-k a=0$. This concludes the proof.

Comment. There are other ways to obtain the divisibility $p \mid f(p)$ for primes $p$, which is all that is needed in this proof. For instance, if $f(p)$ were not divisible by $p$ then the arithmetic progression $p^{2}+b f(p)$ would attain prime values for infinitely many $b$ by Dirichlet's Theorem: hence, for these pairs p , b , we would have $p+f(b)=p^{2}+b f(p)$. Substituting $a \mapsto b$ and $b \mapsto p$ in $(*)$ then shows that $\left(f(p)^{2}-p^{2}\right)(p-1)$ is divisible by $b+f(p)$ and hence vanishes, which is impossible since $p \nmid f(p)$ by assumption.

Solution 2. First, we substitute $b=1$ in (*) and rearrange to find that

$$
\frac{f(a)+f(1)^{2}}{a+f(1)}=f(1)-a+\frac{a^{2}+f(a)}{a+f(1)}
$$

is a positive integer for sufficiently large $a$. Since $f(a) \leqslant a f(1)$, for all sufficiently large $a$, it follows that $\frac{f(a)+f(1)^{2}}{a+f(1)} \leqslant f(1)$ also and hence there is a positive integer $k$ such that $\frac{f(a)+f(1)^{2}}{a+f(1)}=k$ for infinitely many values of $a$. In other words,

$$
f(a)=k a+f(1) \cdot(k-f(1))
$$

for infinitely many $a$.
Fixing an arbitrary choice of $a$ in (*), we have that

$$
\frac{a^{2}+b f(a)}{a+k b+f(1) \cdot(k-f(1))}
$$

is an integer for infinitely many $b$ (the same $b$ as above, maybe with finitely many exceptions). On the other hand, for $b$ taken sufficiently large, this quantity becomes arbitrarily close to $\frac{f(a)}{k}$; this is only possible if $\frac{f(a)}{k}$ is an integer and

$$
\frac{a^{2}+b f(a)}{a+k b+f(1) \cdot(k-f(1))}=\frac{f(a)}{k}
$$

for infinitely many $b$. This rearranges to

$$
\begin{equation*}
\frac{f(a)}{k} \cdot(a+f(1) \cdot(k-f(1)))=a^{2} . \tag{**}
\end{equation*}
$$

Hence $a^{2}$ is divisible by $a+f(1) \cdot(k-f(1))$, and hence so is $f(1)^{2}(k-f(1))^{2}$. The only way this can occur for all $a$ is if $k=f(1)$, in which case ( $* *$ ) provides that $f(a)=k a$ for all $a$, as desired.

Solution 3. Fix any two distinct positive integers $a$ and $b$. From (*) it follows that the two integers

$$
\left(a^{2}+c f(a)\right) \cdot(b+f(c)) \text { and }\left(b^{2}+c f(b)\right) \cdot(a+f(c))
$$

are both multiples of $(a+f(c)) \cdot(b+f(c))$ for all sufficiently large $c$. Taking an appropriate linear combination to eliminate the $c f(c)$ term, we find after expanding out that the integer

$$
\left[a^{2} f(b)-b^{2} f(a)\right] \cdot f(c)+[(b-a) f(a) f(b)] \cdot c+[a b(a f(b)-b f(a))]
$$

is also a multiple of $(a+f(c)) \cdot(b+f(c))$.
But as $c$ varies, $(\dagger)$ is bounded above by a positive multiple of $c$ while $(a+f(c)) \cdot(b+f(c))$ is bounded below by a positive multiple of $c^{2}$. The only way that such a divisibility can hold is if in fact

$$
\left[a^{2} f(b)-b^{2} f(a)\right] \cdot f(c)+[(b-a) f(a) f(b)] \cdot c+[a b(a f(b)-b f(a))]=0
$$

for sufficiently large $c$. Since the coefficient of $c$ in this linear relation is nonzero, it follows that there are constants $k, \ell$ such that $f(c)=k c+\ell$ for all sufficiently large $c$; the constants $k$ and $\ell$ are necessarily integers.

The value of $\ell$ satisfies

$$
\left[a^{2} f(b)-b^{2} f(a)\right] \cdot \ell+[a b(a f(b)-b f(a))]=0
$$

and hence $b \mid \ell a^{2} f(b)$ for all $a$ and $b$. Taking $b$ sufficiently large so that $f(b)=k b+\ell$, we thus have that $b \mid \ell^{2} a^{2}$ for all sufficiently large $b$; this implies that $\ell=0$. From ( $\dagger \dagger \dagger$ ) it then follows that $\frac{f(a)}{a}=\frac{f(b)}{b}$ for all $a \neq b$, so that there is a constant $k$ such that $f(a)=k a$ for all $a(k$ is equal to the constant defined earlier).

Solution 4. Let $\Gamma$ denote the set of all points $(a, f(a))$, so that $\Gamma$ is an infinite subset of the upper-right quadrant of the plane. For a point $A=(a, f(a))$ in $\Gamma$, we define a point $A^{\prime}=\left(-f(a),-f(a)^{2} / a\right)$ in the lower-left quadrant of the plane, and let $\Gamma^{\prime}$ denote the set of all such points $A^{\prime}$.


Claim. For any point $A \in \Gamma$, the set $\Gamma$ is contained in finitely many lines through the point $A^{\prime}$. Proof. Let $A=(a, f(a))$. The functional equation (with $a$ and $b$ interchanged) can be rewritten as $b+f(a) \mid a f(b)-b f(a)$, so that all but finitely many points in $\Gamma$ are contained in one of the lines with equation

$$
a y-f(a) x=m(x+f(a))
$$

for $m$ an integer. Geometrically, these are the lines through $A^{\prime}=\left(-f(a),-f(a)^{2} / a\right)$ with gradient $\frac{f(a)+m}{a}$. Since $\Gamma$ is contained, with finitely many exceptions, in the region $0 \leqslant y \leqslant$ $f(1) \cdot x$ and the point $A^{\prime}$ lies strictly in the lower-left quadrant of the plane, there are only finitely many values of $m$ for which this line meets $\Gamma$. This concludes the proof of the claim.

Now consider any distinct points $A, B \in \Gamma$. It is clear that $A^{\prime}$ and $B^{\prime}$ are distinct. A line through $A^{\prime}$ and a line through $B^{\prime}$ only meet in more than one point if these two lines are equal to the line $A^{\prime} B^{\prime}$. It then follows from the above claim that the line $A^{\prime} B^{\prime}$ must contain all but finitely many points of $\Gamma$. If $C$ is another point of $\Gamma$, then the line $A^{\prime} C^{\prime}$ also passes through all but finitely many points of $\Gamma$, which is only possible if $A^{\prime} C^{\prime}=A^{\prime} B^{\prime}$.

We have thus seen that there is a line $\ell$ passing through all points of $\Gamma^{\prime}$ and through all but finitely many points of $\Gamma$. We claim that this line passes through the origin $O$ and passes through every point of $\Gamma$. To see this, note that by construction $A, O, A^{\prime}$ are collinear for every point $A \in \Gamma$. Since $\ell=A A^{\prime}$ for all but finitely many points $A \in \Gamma$, it thus follows that $O \in \ell$. Thus any $A \in \Gamma$ lies on the line $\ell=A^{\prime} O$.

Since $\Gamma$ is contained in a line through $O$, it follows that there is a real constant $k$ (the gradient of $\ell$ ) such that $f(a)=k a$ for all $a$. The number $k$ is, of course, a positive integer.

Comment. Without the $a+b>C$ condition, this problem is approachable by much more naive methods. For instance, using the given divisibility for $a, b \in\{1,2,3\}$ one can prove by a somewhat tedious case-check that $f(2)=2 f(1)$ and $f(3)=3 f(1)$; this then forms the basis of an induction establishing that $f(n)=n f(1)$ for all $n$.

N5. Let $a$ be a positive integer. We say that a positive integer $b$ is $a$-good if $\binom{a n}{b}-1$ is divisible by $a n+1$ for all positive integers $n$ with $a n \geqslant b$. Suppose $b$ is a positive integer such that $b$ is $a$-good, but $b+2$ is not $a$-good. Prove that $b+1$ is prime.
(Netherlands)
Solution 1. For $p$ a prime and $n$ a nonzero integer, we write $v_{p}(n)$ for the $p$-adic valuation of $n$ : the largest integer $t$ such that $p^{t} \mid n$.

We first show that $b$ is $a$-good if and only if $b$ is even, and $p \mid a$ for all primes $p \leqslant b$.
To start with, the condition that $a n+1 \left\lvert\,\binom{ a n}{b}-1\right.$ can be rewritten as saying that

$$
\begin{equation*}
\frac{a n(a n-1) \cdots(a n-b+1)}{b!} \equiv 1 \quad(\bmod a n+1) \tag{1}
\end{equation*}
$$

Suppose, on the one hand, there is a prime $p \leqslant b$ with $p \nmid a$. Take $t=v_{p}(b!)$. Then there exist positive integers $c$ such that $a c \equiv 1\left(\bmod p^{t+1}\right)$. If we take $c$ big enough, and then take $n=(p-1) c$, then $a n=a(p-1) c \equiv p-1\left(\bmod p^{t+1}\right)$ and $a n \geqslant b$. Since $p \leqslant b$, one of the terms of the numerator $a n(a n-1) \cdots(a n-b+1)$ is $a n-p+1$, which is divisible by $p^{t+1}$. Hence the $p$-adic valuation of the numerator is at least $t+1$, but that of the denominator is exactly $t$. This means that $p \left\lvert\,\binom{ a n}{b}\right.$, so $p \nmid\binom{a n}{b}-1$. As $p \mid a n+1$, we get that $a n+1 \nmid\binom{a n}{b}$, so $b$ is not $a$-good.

On the other hand, if for all primes $p \leqslant b$ we have $p \mid a$, then every factor of $b$ ! is coprime to $a n+1$, and hence invertible modulo $a n+1$ : hence $b$ ! is also invertible modulo $a n+1$. Then equation (1) reduces to:

$$
a n(a n-1) \cdots(a n-b+1) \equiv b!\quad(\bmod a n+1)
$$

However, we can rewrite the left-hand side as follows:

$$
a n(a n-1) \cdots(a n-b+1) \equiv(-1)(-2) \cdots(-b) \equiv(-1)^{b} b!\quad(\bmod a n+1)
$$

Provided that $a n>1$, if $b$ is even we deduce $(-1)^{b} b!\equiv b$ ! as needed. On the other hand, if $b$ is odd, and we take an $+1>2(b!)$, then we will not have $(-1)^{b} b!\equiv b$ !, so $b$ is not $a$-good. This completes the claim.

To conclude from here, suppose that $b$ is $a$-good, but $b+2$ is not. Then $b$ is even, and $p \mid a$ for all primes $p \leqslant b$, but there is a prime $q \leqslant b+2$ for which $q \nmid a$ : so $q=b+1$ or $q=b+2$. We cannot have $q=b+2$, as that is even too, so we have $q=b+1$ : in other words, $b+1$ is prime.

Solution 2. We show only half of the claim of the previous solution: we show that if $b$ is $a$-good, then $p \mid a$ for all primes $p \leqslant b$. We do this with Lucas' theorem.

Suppose that we have $p \leqslant b$ with $p \nmid a$. Then consider the expansion of $b$ in base $p$; there will be some digit (not the final digit) which is nonzero, because $p \leqslant b$. Suppose it is the $p^{t}$ digit for $t \geqslant 1$.

Now, as $n$ varies over the integers, an +1 runs over all residue classes modulo $p^{t+1}$; in particular, there is a choice of $n$ (with $a n>b$ ) such that the $p^{0}$ digit of $a n$ is $p-1$ (so $p \mid a n+1)$ and the $p^{t}$ digit of $a n$ is 0 . Consequently, $p \mid a n+1$ but $p \left\lvert\,\binom{ a n}{b}\right.$ (by Lucas' theorem) so $p \nmid\binom{a n}{b}-1$. Thus $b$ is not $a$-good.

Now we show directly that if $b$ is $a$-good but $b+2$ fails to be so, then there must be a prime dividing $a n+1$ for some $n$, which also divides $(b+1)(b+2)$. Indeed, the ratio between $\binom{a n}{b+2}$ and $\binom{a n}{b}$ is $(b+1)(b+2) /(a n-b)(a n-b-1)$. We know that there must be a choice of $a n+1$ such that the former binomial coefficient is 1 modulo $a n+1$ but the latter is not, which means that the given ratio must not be $1 \bmod a n+1$. If $b+1$ and $b+2$ are both coprime to $a n+1$ then
the ratio $i$ is 1 , so that must not be the case. In particular, as any prime less than $b$ divides $a$, it must be the case that either $b+1$ or $b+2$ is prime.

However, we can observe that $b$ must be even by insisting that $a n+1$ is prime (which is possible by Dirichlet's theorem) and hence $\binom{a n}{b} \equiv(-1)^{b}=1$. Thus $b+2$ cannot be prime, so $b+1$ must be prime.

N6. Let $H=\left\{\lfloor i \sqrt{2}\rfloor: i \in \mathbb{Z}_{>0}\right\}=\{1,2,4,5,7, \ldots\}$, and let $n$ be a positive integer. Prove that there exists a constant $C$ such that, if $A \subset\{1,2, \ldots, n\}$ satisfies $|A| \geqslant C \sqrt{n}$, then there exist $a, b \in A$ such that $a-b \in H$. (Here $\mathbb{Z}_{>0}$ is the set of positive integers, and $\lfloor z\rfloor$ denotes the greatest integer less than or equal to $z$.)
(Brazil)
Common remarks. In all solutions, we will assume that $A$ is a set such that $\{a-b: a, b \in A\}$ is disjoint from $H$, and prove that $|A|<C \sqrt{n}$.

Solution 1. First, observe that if $n$ is a positive integer, then $n \in H$ exactly when

$$
\begin{equation*}
\left\{\frac{n}{\sqrt{2}}\right\}>1-\frac{1}{\sqrt{2}} . \tag{1}
\end{equation*}
$$

To see why, observe that $n \in H$ if and only if $0<i \sqrt{2}-n<1$ for some $i \in \mathbb{Z}_{>0}$. In other words, $0<i-n / \sqrt{2}<1 / \sqrt{2}$, which is equivalent to (1).

Now, write $A=\left\{a_{1}<a_{2}<\cdots<a_{k}\right\}$, where $k=|A|$. Observe that the set of differences is not altered by shifting $A$, so we may assume that $A \subseteq\{0,1, \ldots, n-1\}$ with $a_{1}=0$.

From (1), we learn that $\left\{a_{i} / \sqrt{2}\right\}<1-1 / \sqrt{2}$ for each $i>1$ since $a_{i}-a_{1} \notin H$. Furthermore, we must have $\left\{a_{i} / \sqrt{2}\right\}<\left\{a_{j} / \sqrt{2}\right\}$ whenever $i<j$; otherwise, we would have

$$
-\left(1-\frac{1}{\sqrt{2}}\right)<\left\{\frac{a_{j}}{\sqrt{2}}\right\}-\left\{\frac{a_{i}}{\sqrt{2}}\right\}<0 .
$$

Since $\left\{\left(a_{j}-a_{i}\right) / \sqrt{2}\right\}=\left\{a_{j} / \sqrt{2}\right\}-\left\{a_{i} / \sqrt{2}\right\}+1$, this implies that $\left\{\left(a_{j}-a_{i}\right) / \sqrt{2}\right\}>1 / \sqrt{2}>$ $1-1 / \sqrt{2}$, contradicting (1).

Now, we have a sequence $0=a_{1}<a_{2}<\cdots<a_{k}<n$, with

$$
0=\left\{\frac{a_{1}}{\sqrt{2}}\right\}<\left\{\frac{a_{2}}{\sqrt{2}}\right\}<\cdots<\left\{\frac{a_{k}}{\sqrt{2}}\right\}<1-\frac{1}{\sqrt{2}} .
$$

We use the following fact: for any $d \in \mathbb{Z}$, we have

$$
\begin{equation*}
\left\{\frac{d}{\sqrt{2}}\right\}>\frac{1}{2 d \sqrt{2}} \tag{2}
\end{equation*}
$$

To see why this is the case, let $h=\lfloor d / \sqrt{2}\rfloor$, so $\{d / \sqrt{2}\}=d / \sqrt{2}-h$. Then

$$
\left\{\frac{d}{\sqrt{2}}\right\}\left(\frac{d}{\sqrt{2}}+h\right)=\frac{d^{2}-2 h^{2}}{2} \geqslant \frac{1}{2},
$$

since the numerator is a positive integer. Because $d / \sqrt{2}+h<2 d / \sqrt{2}$, inequality (2) follows.
Let $d_{i}=a_{i+1}-a_{i}$, for $1 \leqslant i<k$. Then $\left\{a_{i+1} / \sqrt{2}\right\}-\left\{a_{i} / \sqrt{2}\right\}=\left\{d_{i} / \sqrt{2}\right\}$, and we have

$$
\begin{equation*}
1-\frac{1}{\sqrt{2}}>\sum_{i}\left\{\frac{d_{i}}{\sqrt{2}}\right\}>\frac{1}{2 \sqrt{2}} \sum_{i} \frac{1}{d_{i}} \geqslant \frac{(k-1)^{2}}{2 \sqrt{2}} \frac{1}{\sum_{i} d_{i}}>\frac{(k-1)^{2}}{2 \sqrt{2}} \cdot \frac{1}{n} . \tag{3}
\end{equation*}
$$

Here, the first inequality holds because $\left\{a_{k} / \sqrt{2}\right\}<1-1 / \sqrt{2}$, the second follows from (2), the third follows from an easy application of the AM-HM inequality (or Cauchy-Schwarz), and the fourth follows from the fact that $\sum_{i} d_{i}=a_{k}<n$.

Rearranging this, we obtain

$$
\sqrt{2 \sqrt{2}-2} \cdot \sqrt{n}>k-1
$$

which provides the required bound on $k$.

Solution 2. Let $\alpha=2+\sqrt{2}$, so $(1 / \alpha)+(1 / \sqrt{2})=1$. Thus, $J=\left\{\lfloor i \alpha\rfloor: i \in \mathbb{Z}_{>0}\right\}$ is the complementary Beatty sequence to $H$ (in other words, $H$ and $J$ are disjoint with $H \cup J=\mathbb{Z}_{>0}$ ). Write $A=\left\{a_{1}<a_{2}<\cdots<a_{k}\right\}$. Suppose that $A$ has no differences in $H$, so all its differences are in $J$ and we can set $a_{i}-a_{1}=\left\lfloor\alpha b_{i}\right\rfloor$ for $b_{i} \in \mathbb{Z}_{>0}$.

For any $j>i$, we have $a_{j}-a_{i}=\left\lfloor\alpha b_{j}\right\rfloor-\left\lfloor\alpha b_{i}\right\rfloor$. Because $a_{j}-a_{i} \in J$, we also have $a_{j}-a_{i}=\lfloor\alpha t\rfloor$ for some positive integer $t$. Thus, $\lfloor\alpha t\rfloor=\left\lfloor\alpha b_{j}\right\rfloor-\left\lfloor\alpha b_{i}\right\rfloor$. The right hand side must equal either $\left\lfloor\alpha\left(b_{j}-b_{i}\right)\right\rfloor$ or $\left\lfloor\alpha\left(b_{j}-b_{i}\right)\right\rfloor-1$, the latter of which is not a member of $J$ as $\alpha>2$. Therefore, $t=b_{j}-b_{i}$ and so we have $\left\lfloor\alpha b_{j}\right\rfloor-\left\lfloor\alpha b_{i}\right\rfloor=\left\lfloor\alpha\left(b_{j}-b_{i}\right)\right\rfloor$.

For $1 \leqslant i<k$ we now put $d_{i}=b_{i+1}-b_{i}$, and we have

$$
\left\lfloor\alpha \sum_{i} d_{i}\right\rfloor=\left\lfloor\alpha b_{k}\right\rfloor=\sum_{i}\left\lfloor\alpha d_{i}\right\rfloor ;
$$

that is, $\sum_{i}\left\{\alpha d_{i}\right\}<1$. We also have

$$
1+\left\lfloor\alpha \sum_{i} d_{i}\right\rfloor=1+a_{k}-a_{1} \leqslant a_{k} \leqslant n
$$

so $\sum_{i} d_{i} \leqslant n / \alpha$.
With the above inequalities, an argument similar to (3) (which uses the fact that $\{\alpha d\}=$ $\{d \sqrt{2}\}>1 /(2 d \sqrt{2})$ for positive integers $d)$ proves that $1>\left((k-1)^{2} /(2 \sqrt{2})\right)(\alpha / n)$, which again rearranges to give

$$
\sqrt{2 \sqrt{2}-2} \cdot \sqrt{n}>k-1
$$

Comment. The use of Beatty sequences in Solution 2 is essentially a way to bypass (1). Both Solutions 1 and 2 use the fact that $\sqrt{2}<2$; the statement in the question would still be true if $\sqrt{2}$ did not have this property (for instance, if it were replaced with $\alpha$ ), but any argument along the lines of Solutions 1 or 2 would be more complicated.

Solution 3. Again, define $J=\mathbb{Z}_{>0} \backslash H$, so all differences between elements of $A$ are in $J$. We start by making the following observation. Suppose we have a set $B \subseteq\{1,2, \ldots, n\}$ such that all of the differences between elements of $B$ are in $H$. Then $|A| \cdot|B| \leqslant 2 n$.

To see why, observe that any two sums of the form $a+b$ with $a \in A, b \in B$ are different; otherwise, we would have $a_{1}+b_{1}=a_{2}+b_{2}$, and so $\left|a_{1}-a_{2}\right|=\left|b_{2}-b_{1}\right|$. However, then the left hand side is in $J$ whereas the right hand side is in $H$. Thus, $\{a+b: a \in A, b \in B\}$ is a set of size $|A| \cdot|B|$ all of whose elements are no greater than $2 n$, yielding the claimed inequality.

With this in mind, it suffices to construct a set $B$, all of whose differences are in $H$ and whose size is at least $C^{\prime} \sqrt{n}$ for some constant $C^{\prime}>0$.

To do so, we will use well-known facts about the negative Pell equation $X^{2}-2 Y^{2}=-1$; in particular, that there are infinitely many solutions and the values of $X$ are given by the recurrence $X_{1}=1, X_{2}=7$ and $X_{m}=6 X_{m-1}-X_{m-2}$. Therefore, we may choose $X$ to be a solution with $\sqrt{n} / 6<X \leqslant \sqrt{n}$.

Now, we claim that we may choose $B=\{X, 2 X, \ldots,\lfloor(1 / 3) \sqrt{n}\rfloor X\}$. Indeed, we have

$$
\left(\frac{X}{\sqrt{2}}-Y\right)\left(\frac{X}{\sqrt{2}}+Y\right)=\frac{-1}{2}
$$

and so

$$
0>\left(\frac{X}{\sqrt{2}}-Y\right) \geqslant \frac{-3}{\sqrt{2 n}}
$$

from which it follows that $\{X / \sqrt{2}\}>1-(3 / \sqrt{2 n})$. Combined with (1), this shows that all differences between elements of $B$ are in $H$.

Comment. Some of the ideas behind Solution 3 may be used to prove that the constant $C=\sqrt{2 \sqrt{2}-2}$ (from Solutions 1 and 2) is optimal, in the sense that there are arbitrarily large values of $n$ and sets $A_{n} \subseteq\{1,2, \ldots, n\}$ of size roughly $C \sqrt{n}$, all of whose differences are contained in $J$.

To see why, choose $X$ to come from a sufficiently large solution to the Pell equation $X^{2}-2 Y^{2}=1$, so $\{X / \sqrt{2}\} \approx 1 /(2 X \sqrt{2})$. In particular, $\{X\},\{2 X\}, \ldots,\{[2 X \sqrt{2}(1-1 / \sqrt{2})\rfloor X\}$ are all less than $1-1 / \sqrt{2}$. Thus, by (1) any positive integer of the form $i X$ for $1 \leqslant i \leqslant\lfloor 2 X \sqrt{2}(1-1 / \sqrt{2})\rfloor$ lies in $J$.

Set $n \approx 2 X^{2} \sqrt{2}(1-1 / \sqrt{2})$. We now have a set $A=\{i X: i \leqslant\lfloor 2 X \sqrt{2}(1-1 / \sqrt{2})\rfloor\}$ containing roughly $2 X \sqrt{2}(1-1 / \sqrt{2})$ elements less than or equal to $n$ such that all of the differences lie in $J$, and we can see that $|A| \approx C \sqrt{n}$ with $C=\sqrt{2 \sqrt{2}-2}$.

Solution 4. As in Solution 3, we will provide a construction of a large set $B \subseteq\{1,2, \ldots, n\}$, all of whose differences are in $H$.

Choose $Y$ to be a solution to the Pell-like equation $X^{2}-2 Y^{2}= \pm 1$; such solutions are given by the recurrence $Y_{1}=1, Y_{2}=2$ and $Y_{m}=2 Y_{m-1}+Y_{m-2}$, and so we can choose $Y$ such that $n /(3 \sqrt{2})<Y \leqslant n / \sqrt{2}$. Furthermore, it is known that for such a $Y$ and for $1 \leqslant x<Y$,

$$
\begin{equation*}
\{x \sqrt{2}\}+\{(Y-x) \sqrt{2}\}=\{Y / \sqrt{2}\} \tag{4}
\end{equation*}
$$

if $X^{2}-2 Y^{2}=1$, and

$$
\begin{equation*}
\{x \sqrt{2}\}+\{(Y-x) \sqrt{2}\}=1+\{Y / \sqrt{2}\} \tag{5}
\end{equation*}
$$

if $X^{2}-2 Y^{2}=-1$. Indeed, this is a statement of the fact that $X / Y$ is a best rational approximation to $\sqrt{2}$, from below in the first case and from above in the second.

Now, consider the sequence $\{\sqrt{2}\},\{2 \sqrt{2}\}, \ldots,\{(Y-1) \sqrt{2}\}$. The Erdős-Szekeres theorem tells us that this sequence has a monotone subsequence with at least $\sqrt{Y-2}+1>\sqrt{Y}$ elements; if that subsequence is decreasing, we may reflect (using (4) or (5)) to ensure that it is increasing. Call the subsequence $\left\{y_{1} \sqrt{2}\right\},\left\{y_{2} \sqrt{2}\right\}, \ldots,\left\{y_{t} \sqrt{2}\right\}$ for $t>\sqrt{Y}$.

Now, set $B=\left\{\left\lfloor y_{i} \sqrt{2}\right\rfloor: 1 \leqslant i \leqslant t\right\}$. We have $\left\lfloor y_{j} \sqrt{2}\right\rfloor-\left\lfloor y_{i} \sqrt{2}\right\rfloor=\left\lfloor\left(y_{j}-y_{i}\right) \sqrt{2}\right\rfloor$ for $i<j$ (because the corresponding inequality for the fractional parts holds by the ordering assumption on the $\left\{y_{i} \sqrt{2}\right\}$ ), which means that all differences between elements of $B$ are indeed in $H$. Since $|B|>\sqrt{Y}>\sqrt{n} / \sqrt{3 \sqrt{2}}$, this is the required set.

Comment. Any solution to this problem will need to use the fact that $\sqrt{2}$ cannot be approximated well by rationals, either directly or implicitly (for example, by using facts about solutions to Pelllike equations). If $\sqrt{2}$ were replaced by a value of $\theta$ with very good rational approximations (from below), then an argument along the lines of Solution 3 would give long arithmetic progressions in $\{[i \theta]: 0 \leqslant i<n\}$ (with initial term 0 ) for certain values of $n$.

N7. Prove that there is a constant $c>0$ and infinitely many positive integers $n$ with the following property: there are infinitely many positive integers that cannot be expressed as the sum of fewer than $c n \log (n)$ pairwise coprime $n^{\text {th }}$ powers.
(Canada)

Solution 1. Suppose, for an integer $n$, that we can find another integer $N$ satisfying the following property:

$$
n \text { is divisible by } \varphi\left(p^{e}\right) \text { for every prime power } p^{e} \text { exactly dividing } N \text {. }
$$

This property ensures that all $n^{\text {th }}$ powers are congruent to 0 or 1 modulo each such prime power $p^{e}$, and hence that any sum of $m$ pairwise coprime $n^{\text {th }}$ powers is congruent to $m$ or $m-1$ modulo $p^{e}$, since at most one of the $n^{\text {th }}$ powers is divisible by $p$. Thus, if $k$ denotes the number of distinct prime factors of $N$, we find by the Chinese Remainder Theorem at most $2^{k} m$ residue classes modulo $N$ which are sums of at most $m$ pairwise coprime $n^{\text {th }}$ powers. In particular, if $N>2^{k} m$ then there are infinitely many positive integers not expressible as a sum of at most $m$ pairwise coprime $n^{\text {th }}$ powers.

It thus suffices to prove that there are arbitrarily large pairs $(n, N)$ of integers satisfying $(\dagger)$ such that

$$
N>c \cdot 2^{k} n \log (n)
$$

for some positive constant $c$.
We construct such pairs as follows. Fix a positive integer $t$ and choose (distinct) prime numbers $p \mid 2^{2^{t-1}}+1$ and $q \mid 2^{2^{t}}+1$; we set $N=p q$. It is well-known that $2^{t} \mid p-1$ and $2^{t+1} \mid q-1$, hence

$$
n=\frac{(p-1)(q-1)}{2^{t}}
$$

is an integer and the pair $(n, N)$ satisfies ( $\dagger$ ).
Estimating the size of $N$ and $n$ is now straightforward. We have

$$
\log _{2}(n) \leqslant 2^{t-1}+2^{t}-t<2^{t+1}<2 \cdot \frac{N}{n}
$$

which rearranges to

$$
N>\frac{1}{8} \cdot 2^{2} n \log _{2}(n)
$$

and so we are done if we choose $c<\frac{1}{8 \log (2)} \approx 0.18$.
Comment 1. The trick in the above solution was to find prime numbers $p$ and $q$ congruent to 1 modulo some $d=2^{t}$ and which are not too large. An alternative way to do this is via Linnik's Theorem, which says that there are absolute constants $b$ and $L>1$ such that for any coprime integers $a$ and $d$, there is a prime congruent to $a$ modulo $d$ and of size $\leqslant b d^{L}$. If we choose some $d$ not divisible by 3 and choose two distinct primes $p, q \leqslant b \cdot(3 d)^{L}$ congruent to 1 modulo $d$ (and, say, distinct modulo 3), then we obtain a pair $(n, N)$ satisfying $(\dagger)$ with $N=p q$ and $n=\frac{(p-1)(q-1)}{d}$. A straightforward computation shows that

$$
N>C n^{1+\frac{1}{2 L-1}}
$$

for some constant $C$, which is in particular larger than any $c \cdot 2^{2} n \log (n)$ for $p$ large. Thus, the statement of the problem is true for any constant $c$. More strongly, the statement of the problem is still true when $c n \log (n)$ is replaced by $n^{1+\delta}$ for a sufficiently small $\delta>0$.

Solution 2, obtaining better bounds. As in the preceding solution, we seek arbitrarily large pairs of integers $n$ and $N$ satisfying ( $\dagger$ ) such that $N>c 2^{k} n \log (n)$.

This time, to construct such pairs, we fix an integer $x \geqslant 4$, set $N$ to be the lowest common multiple of $1,2, \ldots, 2 x$, and set $n$ to be twice the lowest common multiple of $1,2, \ldots, x$. The pair $(n, N)$ does indeed satisfy the condition, since if $p^{e}$ is a prime power divisor of $N$ then $\frac{\varphi\left(p^{e}\right)}{2} \leqslant x$ is a factor of $\frac{n}{2}=\operatorname{lcm}_{r \leqslant x}(r)$.

Now $2 N / n$ is the product of all primes having a power lying in the interval $(x, 2 x]$, and hence $2 N / n>x^{\pi(2 x)-\pi(x)}$. Thus for sufficiently large $x$ we have

$$
\log \left(\frac{2 N}{2^{\pi(2 x)} n}\right)>(\pi(2 x)-\pi(x)) \log (x)-\log (2) \pi(2 x) \sim x
$$

using the Prime Number Theorem $\pi(t) \sim t / \log (t)$.
On the other hand, $n$ is a product of at most $\pi(x)$ prime powers less than or equal to $x$, and so we have the upper bound

$$
\log (n) \leqslant \pi(x) \log (x) \sim x
$$

again by the Prime Number Theorem. Combined with the above inequality, we find that for any $\epsilon>0$, the inequality

$$
\log \left(\frac{N}{2^{\pi(2 x)} n}\right)>(1-\epsilon) \log (n)
$$

holds for sufficiently large $x$. Rearranging this shows that

$$
N>2^{\pi(2 x)} n^{2-\epsilon}>2^{\pi(2 x)} n \log (n)
$$

for all sufficiently large $x$ and we are done.
Comment 2. The stronger bound $N>2^{\pi(2 x)} n^{2-\epsilon}$ obtained in the above proof of course shows that infinitely many positive integers cannot be written as a sum of at most $n^{2-\epsilon}$ pairwise coprime $n^{\text {th }}$ powers.

By refining the method in Solution 2, these bounds can be improved further to show that infinitely many positive integers cannot be written as a sum of at most $n^{\alpha}$ pairwise coprime $n^{\text {th }}$ powers for any positive $\alpha>0$. To do this, one fixes a positive integer $d$, sets $N$ equal to the product of the primes at most $d x$ which are congruent to 1 modulo $d$, and $n=d \mathrm{lcm}_{r \leqslant x}(r)$. It follows as in Solution 2 that $(n, N)$ satisfies $(\dagger)$.

Now the Prime Number Theorem in arithmetic progressions provides the estimates $\log (N) \sim \frac{d}{\varphi(d)} x$, $\log (n) \sim x$ and $\pi(d x) \sim \frac{d x}{\log (x)}$ for any fixed $d$. Combining these provides a bound

$$
N>2^{\pi(d x)} n^{d / \varphi(d)-\epsilon}
$$

for any positive $\epsilon$, valid for $x$ sufficiently large. Since the ratio $\frac{d}{\varphi(d)}$ can be made arbitrarily large by a judicious choice of $d$, we obtain the $n^{\alpha}$ bound claimed.

Comment 3. While big results from analytic number theory such as the Prime Number Theorem or Linnik's Theorem certainly can be used in this problem, they do not seem to substantially simplify matters: all known solutions involve first reducing to condition ( $\dagger$ ), and even then analytic results do not make it clear how to proceed. For this reason, we regard this problem as suitable for the IMO.

Rather than simplifying the problem, what nonelementary results from analytic number theory allow one to achieve is a strengthening of the main bound, typically replacing the $n \log (n)$ growth with a power $n^{1+\delta}$. However, we believe that such stronger bounds are unlikely to be found by students in the exam.

The strongest bound we know how to achieve using purely elementary methods is a bound of the form $N>2^{k} n \log (n)^{M}$ for any positive integer $M$. This is achieved by a variant of the argument in Solution 1, choosing primes $p_{0}, \ldots, p_{M}$ with $p_{i} \mid 2^{2^{t+i-1}}+1$ and setting $N=\prod_{i} p_{i}$ and $n=$ $2^{-t M} \prod_{i}\left(p_{i}-1\right)$.

N8. Let $a$ and $b$ be two positive integers. Prove that the integer

$$
a^{2}+\left\lceil\frac{4 a^{2}}{b}\right\rceil
$$

is not a square. (Here $\lceil z\rceil$ denotes the least integer greater than or equal to $z$.)
(Russia)
Solution 1. Arguing indirectly, assume that

$$
a^{2}+\left\lceil\frac{4 a^{2}}{b}\right\rceil=(a+k)^{2}, \quad \text { or } \quad\left\lceil\frac{(2 a)^{2}}{b}\right\rceil=(2 a+k) k
$$

Clearly, $k \geqslant 1$. In other words, the equation

$$
\begin{equation*}
\left\lceil\frac{c^{2}}{b}\right\rceil=(c+k) k \tag{1}
\end{equation*}
$$

has a positive integer solution $(c, k)$, with an even value of $c$.
Choose a positive integer solution of (1) with minimal possible value of $k$, without regard to the parity of $c$. From

$$
\frac{c^{2}}{b}>\left\lceil\frac{c^{2}}{b}\right\rceil-1=c k+k^{2}-1 \geqslant c k
$$

and

$$
\frac{(c-k)(c+k)}{b}<\frac{c^{2}}{b} \leqslant\left\lceil\frac{c^{2}}{b}\right\rceil=(c+k) k
$$

it can be seen that $c>b k>c-k$, so

$$
c=k b+r \quad \text { with some } 0<r<k .
$$

By substituting this in (1) we get

$$
\left\lceil\frac{c^{2}}{b}\right\rceil=\left\lceil\frac{(b k+r)^{2}}{b}\right\rceil=k^{2} b+2 k r+\left\lceil\frac{r^{2}}{b}\right\rceil
$$

and

$$
(c+k) k=(k b+r+k) k=k^{2} b+2 k r+k(k-r),
$$

so

$$
\begin{equation*}
\left\lceil\frac{r^{2}}{b}\right\rceil=k(k-r) \tag{2}
\end{equation*}
$$

Notice that relation (2) provides another positive integer solution of (1), namely $c^{\prime}=r$ and $k^{\prime}=k-r$, with $c^{\prime}>0$ and $0<k^{\prime}<k$. That contradicts the minimality of $k$, and hence finishes the solution.

Solution 2. Suppose that

$$
a^{2}+\left\lceil\frac{4 a^{2}}{b}\right\rceil=c^{2}
$$

with some positive integer $c>a$, so

$$
\begin{align*}
& c^{2}-1<a^{2}+\frac{4 a^{2}}{b} \leqslant c^{2} \\
& 0 \leqslant c^{2} b-a^{2}(b+4)<b \tag{3}
\end{align*}
$$

Let $d=c^{2} b-a^{2}(b+4), x=c+a$ and $y=c-a$; then we have $c=\frac{x+y}{2}$ and $a=\frac{x-y}{2}$, and (3) can be re-written as follows:

$$
\begin{align*}
\left(\frac{x+y}{2}\right)^{2} b-\left(\frac{x-y}{2}\right)^{2}(b+4) & =d \\
x^{2}-(b+2) x y+y^{2}+d=0 ; \quad 0 & \leqslant d<b . \tag{4}
\end{align*}
$$

So, by the indirect assumption, the equation (4) has some positive integer solution $(x, y)$.
Fix $b$ and $d$, and take a pair $(x, y)$ of positive integers, satisfying (4), such that $x+y$ is minimal. By the symmetry in (4) we may assume that $x \geqslant y \geqslant 1$.

Now we perform a usual "Vieta jump". Consider (4) as a quadratic equation in variable $x$, and let $z$ be its second root. By the Vieta formulas,

$$
x+z=(b+2) y, \quad \text { and } \quad z x=y^{2}+d,
$$

so

$$
z=(b+2) y-x=\frac{y^{2}+d}{x} .
$$

The first formula shows that $z$ is an integer, and by the second formula $z$ is positive. Hence $(z, y)$ is another positive integer solution of (4). From

$$
\begin{aligned}
(x-1)(z-1) & =x z-(x+z)+1=\left(y^{2}+d\right)-(b+2) y+1 \\
& <\left(y^{2}+b\right)-(b+2) y+1=(y-1)^{2}-b(y-1) \leqslant(y-1)^{2} \leqslant(x-1)^{2}
\end{aligned}
$$

we can see that $z<x$ and therefore $z+y<x+y$. But this contradicts the minimality of $x+y$ among the positive integer solutions of (4).

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[^0]:    *The name Dirichlet interval is chosen for the reason that $g$ theoretically might act similarly to the Dirichlet function on this interval.

