DOI: 10.2478/udt-2022-0012
Unif. Distrib. Theory 17 (2022), no.2, 51-76

# CONSECUTIVE RATIOS IN SECOND-ORDER LINEAR RECURRENCE SEQUENCES 

Daniel Berend ${ }^{1}$ - Rishi Kumar ${ }^{2}$<br>${ }^{1}$ Ben-Gurion University of the Negev, Beer Sheva, ISRAEL<br>${ }^{2}$ Ben-Gurion University of the Negev, Beer Sheva, ISRAEL


#### Abstract

Let $\left(a_{n}\right)_{n=0}^{\infty}$ be a second-order linear recurrence sequence with constant coefficient. We study the limit points and asymptotic distribution of the sequence of consecutive ratios $a_{n+1} / a_{n}$.


Communicated by Artūras Dubickas

## 1. Introduction

The Fibonacci sequence $\left(F_{n}\right)_{n=0}^{\infty}$ is defined by $F_{0}=1, F_{1}=1$, and

$$
\begin{equation*}
F_{n}=F_{n-1}+F_{n-2}, \quad n>1 . \tag{1}
\end{equation*}
$$

The limit of the ratios of consecutive terms of $\left(F_{n}\right)_{n=0}^{\infty}$ is well known to be

$$
\lim _{n \rightarrow \infty} \frac{F_{n+1}}{F_{n}}=\phi
$$

where $\phi=\frac{1+\sqrt{5}}{2}=1.618 \ldots$ is the golden mean [22, p.240].
In general, a sequence of complex numbers $\left(a_{n}\right)_{n=0}^{\infty}$ is a linear recurrence sequence of order $k$ with constant coefficients if it satisfies:

$$
\begin{equation*}
a_{n}=c_{k-1} a_{n-1}+\cdots+c_{0} a_{n-k}, \quad n \geq k,\left(c_{0}, \ldots, c_{k-1} \in \mathbf{C}\right) \tag{2}
\end{equation*}
$$

[^0]
## DANIEL BEREND - RISHI KUMAR

The polynomial

$$
p(\lambda)=\lambda^{k}-c_{k-1} \lambda^{k-1}-\cdots-c_{0}
$$

is the characteristic polynomial of $\left(a_{n}\right)_{n=0}^{\infty}$. Let $\lambda_{1}, \ldots, \lambda_{h}$ be the roots of $p$, with respective multiplicities $k_{1}, \ldots, k_{h}$. The general term $a_{n}$ may be written explicitly in the form

$$
\begin{equation*}
a_{n}=\sum_{i=1}^{h} \sum_{j=1}^{k_{i}} c_{i, j} n^{j-1} \lambda_{i}^{n}, \quad n \in \mathbf{N}, \tag{3}
\end{equation*}
$$

where the coefficients $c_{i, j}$ are complex numbers, uniquely determined by $a_{0}, a_{1}$, $, \ldots, a_{k-1}$ (see [14, Theorem 3.6]). Usually, there is one term on the right-hand side of (3) that dominates all others. In fact, order the roots $\lambda_{i}$ so that $\left|\lambda_{1}\right| \geq$ $\left|\lambda_{2}\right| \geq \cdots \geq\left|\lambda_{h}\right|$; between roots $\lambda_{i}$ and $\lambda_{j}$ of the same modulus, $\lambda_{i}$ precedes $\lambda_{j}$ if $k_{i}>k_{j}$. To avoid trivialities, we assume that $c_{i, k_{i}} \neq 0$ for each $i$. If $\left|\lambda_{1}\right|>\left|\lambda_{2}\right|$, or $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$ and $k_{1}>k_{2}$, then the term $c_{1, k_{1}} n^{k_{1}-1} \lambda_{1}^{n}$ is much larger in absolute value than all other terms for large $n$. In this case, for large $n$ we have

$$
\frac{a_{n+1}}{a_{n}} \approx \frac{c_{1, k_{1}}(n+1)^{k_{1}-1} \lambda_{1}^{n+1}}{c_{1, k_{1}} n^{k_{1}-1} \lambda_{1}^{n}},
$$

and therefore $a_{n+1} / a_{n} \xrightarrow[n \rightarrow \infty]{ } \lambda_{1}$. (Here and later, if finitely many $a_{n}$-s vanish, we consider the ratios $a_{n+1} / a_{n}$ only for sufficiently large $n$.) In particular, if $\left|\lambda_{1}\right|>\left|\lambda_{2}\right|>\cdots>\left|\lambda_{h}\right|$, then, by omitting 0 terms from the right-hand side of (3), we see that the consecutive ratios $a_{n+1} / a_{n}$ converge to one of the $\lambda_{i}$ as $n \rightarrow \infty$. (This is a special case of a result of Poincaré [27].) On the other hand, if there exist two distinct roots of $p$ with the same modulus, then it is always possible to find initial conditions so that $\lim _{n \rightarrow \infty} a_{n+1} / a_{n}$ does not exist [7].

The special case where $\left(a_{n}\right)_{n=0}^{\infty}$ is a sequence of integers was studied in 1518, 20, 21. Suppose the roots of its characteristic polynomial are distinct and satisfy $\left|\lambda_{1}\right|>\left|\lambda_{2}\right| \geq \cdots \geq\left|\lambda_{h}\right|$. We may write $a_{n}$ in the form

$$
a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}+\cdots+c_{h} \lambda_{h}^{n}, \quad n \geq 0
$$

where $c_{1}, c_{2}, \ldots, c_{h}$ are algebraic numbers and assume that $c_{1} \neq 0$. The sequence of ratios $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$ converges to $\lambda_{1}$. In [15, 17, 18, the rate of convergence of the ratios $a_{n+1} / a_{n}$ to this limit was studied. If $\left(a_{n}\right)_{n=0}^{\infty}$ is of order 2 and $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$, it was shown that $\left|\lambda_{1}\right|$ is a partial limit of the sequence of ratios, and the distances between the terms of the sequence and this partial limit were discussed [16, 18, 21].

Following [7], we call $\lim _{n \rightarrow \infty} a_{n+1} / a_{n}$, if it exists, the Kepler limit of $\left(a_{n}\right)_{n=0}^{\infty}$. The Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is the set of limit points of the sequence $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$. In this paper, by a linear recurrence sequence we mean a linear recurrence sequence with constant coefficients. To avoid trivialities, we will always assume that the sequence is not identically 0 .

Bagdasar, Hedderwick, and Popa [2] discussed the Kepler set of second-order linear recurrence sequences. According to the discussion above, unless the roots $\lambda_{1}$ and $\lambda_{2}$ of the characteristic polynomial $p$ are of equal moduli, the Kepler set reduces to a single point. For $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$, they noted that the Kepler set may be a finite set or a circle in the complex plane. We will notice that the Kepler set may be also a line. Our first goal will be to understand the exact dependence of the Kepler set on the parameters in the representation $a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}$. We will also characterize the lines and circles that arise as Kepler sets.

For general linear recurrence sequences, in addition to the topological information given by the Kepler set, one may ask about the distribution of the sequence $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$ in the complex plane. For example, suppose that the Kepler set is a circle. Does the sequence spend roughly the same time in equal arcs of this circle? Taking any example, one is readily convinced that this is not the case (see, for example Figure 1 (B) in the sequel). Our main goal in this paper is to understand how the sequence of consecutive ratios is distributed for any linear recurrence sequence of order 2 . Moreover, we will characterize the family of distributions on the complex plane arising this way.

The distribution modulo 1 of the sequence of consecutive ratios was studied in several papers [8, 19 ] when $\left(a_{n}\right)_{n=0}^{\infty}$ is a real-valued recurrence sequence. Kiss and Tichy [19] studied the distribution for order-2 real-valued sequences $\left(a_{n}\right)_{n=0}^{\infty}$. Under suitable conditions, they determined the asymptotic distribution function of the sequence $\left(a_{n+1} / a_{n}\right)$ and gave an estimate of the error term. The asymptotic distribution function of the sequence $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$, for linear recurrence sequence of any order $k \geq 2$, was discussed by Goldstern, Tichy, and Turnwald [8].

The sequence of consecutive ratios is easily seen to be the orbit if some point in the complex plane under a certain Möbius transformation. We will discuss the ergodic-theoretical properties of this transformation and derive some properties of the sequence of ratios.

In Section 2, we state our main results. Section 3 introduces an extended family of Cauchy distributions in the complex plane and studies their behaviour under Möbius transformations. In Section 4, we discuss the intuition behind the main results. Section 5 presents the proofs.

## DANIEL BEREND — RISHI KUMAR

## 2. Main results

Let $\left(a_{n}\right)_{n=0}^{\infty}$ be a second-order linear recurrence sequence

$$
\begin{equation*}
a_{n}=c a_{n-1}+d a_{n-2}, \quad n \geq 2 \tag{4}
\end{equation*}
$$

with some initial values $a_{0}, a_{1}$, where $c, d$ are fixed complex numbers (with some restrictions listed below, designed to avoid trivialities). We have

$$
\begin{equation*}
\frac{a_{n+2}}{a_{n+1}}=c+\frac{d}{a_{n+1} / a_{n}} . \tag{5}
\end{equation*}
$$

Denote $r_{n}=a_{n+1} / a_{n}$ for $n \geq 0$. By (5), the sequence $\left(r_{n}\right)_{n=1}^{\infty}$ satisfies the recurrence

$$
r_{n+1}=c+d / r_{n} .
$$

Consider the Möbius transformation on the extended complex plane $\mathbf{C}_{\infty}$, defined by

$$
S(z)=c+d / z, \quad z \in \mathbf{C}_{\infty}
$$

In terms of $S$, the Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is the set of limit points of the sequence $\left(S^{n}\left(a_{1} / a_{0}\right)\right)_{n=0}^{\infty}$. (If $a_{n}=0$ for some $n$, we take $a_{n+1} / a_{n}=\infty$. The assumptions below will guarantee that we cannot have $a_{n}=a_{n+1}=0$.) In particular, the Kepler set is $S$-invariant.

Rewrite $a_{n}$ explicitly

$$
\begin{equation*}
a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}, \quad n=0,1,2, \ldots \tag{6}
\end{equation*}
$$

where $\lambda_{1}$ and $\lambda_{2}$ are the roots of the characteristic polynomial $p$ and $c_{1}, c_{2}$ are complex numbers. As explained above, unless $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$, the sequence has a Kepler limit. Thus, we will assume throughout that $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$ (but $\lambda_{1} \neq \lambda_{2}$ ) and that $c_{1}, c_{2} \neq 0$. Moreover, if $\lambda_{2} / \lambda_{1}$ is a root of unity, then the sequence of consecutive ratios is periodic; we will exclude this trivial case.

Lemma 2.1. Let $\lambda_{1}$ and $\lambda_{2}$ be the roots of the characteristic polynomial of $\left(a_{n}\right)_{n=0}^{\infty}$ given in (4). If $\left|\lambda_{1}\right|=\left|\lambda_{1}\right|$ and $\lambda_{2} / \lambda_{1}$ is not a root of unity, then $d<-c^{2} / 4$. Moreover, $a_{n}=0$ for at most one value of $n$.

Denote by $O(C, R)$ the circle of radius $R$, centered at $C$. The following theorem identifies the infinite Kepler sets of order 2 linear recurrence sequences.

## Theorem 2.2.

1. Let $\left(a_{n}\right)_{n=0}^{\infty}$ be a second-order linear recurrence sequence, given by $a_{n}=$ $c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}$ for some non-zero complex numbers $c_{1}, c_{2}, \lambda_{1}, \lambda_{2}$ with $\left|\lambda_{1}\right|=$ $\left|\lambda_{2}\right|>0$, where $\lambda_{2} / \lambda_{1}$ is not a root of unity. Then:
(i) If $\left|c_{1}\right|=\left|c_{2}\right|$, then the Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is the line passing through the origin and the point $\left(\lambda_{1}+\lambda_{2}\right) / 2$.

## CONSECUTIVE RATIOS IN RECURRENCE SEQUENCES

(ii) If $\left|c_{1}\right| \neq\left|c_{2}\right|$, then the Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is the circle $O(C, R)$ with

$$
C=\frac{\left|c_{1}\right|^{2} \lambda_{1}-\left|c_{2}\right|^{2} \lambda_{2}}{\left|c_{1}\right|^{2}-\left|c_{2}\right|^{2}}
$$

and

$$
R=\frac{\left|c_{1}\right|\left|c_{2}\right|\left|\lambda_{1}-\lambda_{2}\right|}{\left|\left|c_{1}\right|^{2}-\left|c_{2}\right|^{2}\right|}
$$

The circle does not include the origin either on its circumference or inside $i t$.
2. Conversely, if $K$ is a line passing through the origin, or a circle not containing the origin either on its circumference or inside, then there exists a second--order linear recurrence sequence whose Kepler set is $K$.

Example. Let $\lambda_{1}=(3+4 i) / 5$ and $\lambda_{2}=(5-12 i) / 13$. For $c_{1}=c_{2}=1$ the Kepler set is the line passing through origin and the point $\left(\lambda_{1}+\lambda_{2}\right) / 2=32 / 65-4 / 65 i$, and for $c_{1}=2, c_{2}=1$, it is the circle $O(131 / 195+268 / 195 i, 28 /(3 \sqrt{65}))$ (see Figure (1).


Figure 1. (A) The Kepler set of $a_{n}=\lambda_{1}^{n}+\lambda_{2}^{n}$, where $\lambda_{1}=(3+4 i) / 5$, $\lambda_{2}=(5-12 i) / 13$. (B) The Kepler set of $a_{n}=2 \lambda_{1}^{n}+\lambda_{2}^{n}$.

Remark 1. The role of the coefficients $c_{1}$ and $c_{2}$ in determining the Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is only via $\left|c_{2} / c_{1}\right|$. Thus, in part 1. (i) of the theorem, the line depends only on $\lambda_{1}$ and $\lambda_{2}$ (as long as $\left|c_{1}\right|=\left|c_{2}\right|$ ). In part 1. (ii) of the theorem, denoting $r=\left|c_{2} / c_{1}\right|$, we may rewrite $C$ and $R$ in a much simpler form:

$$
\begin{equation*}
C=\frac{\lambda_{1}-r^{2} \lambda_{2}}{1-r^{2}}, \quad R=\frac{r\left|\lambda_{1}-\lambda_{2}\right|}{\left|1-r^{2}\right|} . \tag{7}
\end{equation*}
$$

## DANIEL BEREND — RISHI KUMAR

Remark 2. If $\lambda_{2} / \lambda_{1}$ is a primitive root of unity of order $m$, then the sequence $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$ is of period $m$, and in particular the Kepler set is of size $m$. It is contained in the line (if $\left|c_{1}\right|=\left|c_{2}\right|$ ) or circle (if $\left|c_{1}\right| \neq\left|c_{2}\right|$ ) specified in Theorem 2.2. (See Figure 2 for two examples of Kepler sets of size 45.)


Figure 2. (A.) The Kepler set of $a_{n}=\lambda_{1}^{n}+\lambda_{2}^{n}$, where $\lambda_{1}=\exp (4 \pi i / 5)$, $\lambda_{2}=\exp (14 \pi i / 9)$. (B) The Kepler set of $a_{n}=2 \lambda_{1}^{n}+\lambda_{2}^{n}$. Both Kepler sets comprise $9 \cdot 5=45$ points.

Theorem 2.2 specifies the "topology" of the sequence $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$. How is the sequence distributed in the Kepler set? To this end, recall the notion of the distribution of a (deterministic) sequence in a topological space. Let $X$ be a locally compact Hausdorff space, $\mathcal{B}$ its Borel $\sigma$-field, and $\left(x_{n}\right)_{n=0}^{\infty}$ a sequence in $X$. The sequence is distributed according to some probability measure $\nu$ on $(X, \mathcal{B})$ if

$$
\lim _{N \rightarrow \infty} \sum_{n=0}^{N-1} f\left(x_{n}\right)=\int_{X} f \mathrm{~d} \nu
$$

for all continuous function $f: X \rightarrow \mathbf{C}$ (see [23, p. 178]). Intuitively, it means that, for a "well-behaved" set $A \in \mathcal{B}$, the asymptotic density of the set $\left\{n: x_{n} \in A\right\}$ is $\nu(A)$.

## Definition 2.3.

Let $\left(a_{n}\right)_{n=0}^{\infty}$ be a sequence in $\mathbf{C}$. The Kepler measure of $\left(a_{n}\right)_{n=0}^{\infty}$ is the probability measure according to which the sequence $\left(a_{n+1} / a_{n}\right)_{n=0}^{\infty}$ is distributed, if any.

We define two families of distributions that will be relevant in the sequel. A complex-valued random variable $Y$ is Cauchy distributed with median $\mu \in \mathbf{C}$, scale $\sigma \in \mathbf{R}_{+}$, and direction $\alpha \in[0,2 \pi]$, and we denote $Y \sim C(\mu, \sigma, \alpha)$ if:
(i) $Y$ is supported on the line $L$ passing through $\mu$ and making an angle $\alpha$ with the positive real axis. (Here, if $\alpha=0$ then $L$ is parallel to the real axis or coincides with it.)
(ii) The density of $Y$ (with respect to arc-length) is

$$
\begin{equation*}
f_{Y}(z)=\frac{1}{\sigma \pi} \cdot \frac{1}{1+\left|\frac{z-\mu}{\sigma}\right|^{2}}, \quad z \in L \tag{8}
\end{equation*}
$$

A complex-valued random variable $Z$, supported on a circle $O(C, R)$, is circular Cauchy distributed with location $C$, scale $R$, and eccentricity $\psi$ if its density function (with respect to arc-length) is given by

$$
\begin{equation*}
f_{Z}(z)=\frac{R}{2 \pi} \cdot \frac{\left|1-|\psi|^{2}\right|}{|z-(C+R \psi)|^{2}}, \quad z \in O(C, R), \quad(\psi \in \mathbf{C},|\psi| \neq \mathbf{1}) \tag{9}
\end{equation*}
$$

We write $Z \sim C^{\star}(C, R, \psi)$ in this case.
Theorem 2.4. Let $\left(a_{n}\right)$ be a second-order linear recurrence sequence, defined by $a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}$ for some non-zero complex numbers $c_{1}, c_{2}, \lambda_{1}, \lambda_{2}$ with $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|>0$, where $\lambda_{2} / \lambda_{1}$ is not a root of unity.

1. If $\left|c_{1}\right|=\left|c_{2}\right|$, then the Kepler measure of $\left(a_{n}\right)_{n=0}^{\infty}$ is $C(\mu, \sigma, \arg (\mu))$, where

$$
\mu=\frac{\lambda_{1}+\lambda_{2}}{2}, \quad \sigma=\left|\frac{\lambda_{2}-\lambda_{1}}{2}\right| .
$$

2. If $\left|c_{1}\right| \neq\left|c_{2}\right|$, then the Kepler measure of $\left(a_{n}\right)_{n=0}^{\infty}$ is $C^{\star}(C, R, \psi)$, where $R$ and $C$ are as in Theorem 2.2 and

$$
\psi=\frac{\left|c_{2}\right|}{\left|c_{1}\right|} \cdot \frac{\|\left. c_{1}\right|^{2}-\left|c_{2}\right|^{2} \mid}{\left|c_{1}\right|^{2}-\left|c_{2}\right|^{2}} \cdot \frac{\lambda_{2}-\lambda_{1}}{\left|\lambda_{2}-\lambda_{1}\right|} .
$$

In the following theorem we identify the collection of all Kepler measures of second-order linear recurrence sequences with any fixed (infinite) Kepler set. It is analogous to Theorem [2.2, 2.

## Theorem 2.5.

1. The Cauchy measure $C(\mu, \sigma, \alpha)$ is the Kepler measure of some second-order linear recurrence sequence if and only if $\alpha=\arg (\mu)$ and

$$
\begin{equation*}
\frac{1}{\pi} \arctan (\sigma /|\mu|) \notin \mathbf{Q} \tag{10}
\end{equation*}
$$

2. The circular Cauchy distribution $C^{\star}(C, R, \psi)$ is the Kepler measure of some second-order linear recurrence sequence if and only if:
(a) $R_{i}-C-$,
(b) $\psi$ lies on the circle $O\left(C^{\prime}, R^{\prime}\right)$, where we set

$$
\begin{gather*}
C^{\prime}=-C / R \quad \text { and } \quad R^{\prime}=\sqrt{(|C| / R)^{2}-1}, \\
\frac{1}{\pi} \arg \left(\frac{C+R / \bar{\psi}}{C+R \psi}\right) \notin \mathbf{Q} \tag{11}
\end{gather*}
$$

The Kepler measure $\nu$ is invariant under the transformation $S$, namely for every measurable subset $A$ of the Kepler set we have $\nu\left(S^{-1} A\right)=\nu(A)$. Thus, we may study the ergodic-theoretical properties of the system. We recall several basic definitions and results from ergodic theory. (See [31] for more details.)

Let $(X, \mathcal{B}, \nu)$ be a probability space and $T: X \rightarrow X$ be measure-preserving. The pointwise ergodic theorem states that

$$
\begin{equation*}
\frac{1}{n} \sum_{k=0}^{n-1} f\left(T^{k} x\right) \underset{n \rightarrow \infty}{\longrightarrow} \bar{f}(x), \quad f \in L_{1}(\nu) \tag{12}
\end{equation*}
$$

for almost all $x \in X$ with respect to $\nu$, for some $T$-invariant function $\bar{f} \in L_{1}(\nu)$, namely a function satisfying $\bar{f} \circ T=\bar{f}$. $T$ is ergodic if, for every $T$-invariant set $E \subseteq \mathcal{B}$, either $\nu(E)=0$ or $\nu(E)=1$. If $T$ is ergodic, the ergodic theorem takes the simpler form

$$
\begin{equation*}
\frac{1}{n} \sum_{k=0}^{n-1} f\left(T^{k} x\right) \xrightarrow[n \rightarrow \infty]{\text { a.e. }} \int f \mathrm{~d} \nu, \quad f \in L_{1}(\nu) \tag{13}
\end{equation*}
$$

Now let $X$ be a compact metric space and $\mathcal{B}$ its Borel $\sigma$-field. Let $T$ be a continuous transformation from $X$ to itself. It is well known that there exist $T$-invariant probability measures on $(X, \mathcal{B})$ (see [13, Theorem 4.1.1]). Let $\nu$ be such a measure and suppose $T$ is ergodic. A point $x_{0} \in X$ is generic if

$$
\begin{equation*}
\frac{1}{n} \sum_{k=0}^{n-1} f\left(T^{k} x_{0}\right) \underset{n \rightarrow \infty}{\longrightarrow} \int f \mathrm{~d} \nu \tag{14}
\end{equation*}
$$

for every continuous function $f: X \rightarrow \mathbf{C}$. A system $(X, T)$ is uniquely ergodic if $T$ admits a unique invariant probability measure. If $(X, T)$ is uniquely ergodic, then every point $x \in X$ is generic.

Proposition 2.6. In the setup of Theorems 2.2. 1. (ii) and 2.4. 2., the transformation $S$ is uniquely ergodic, and the unique $S$-invariant probability measure
is the Kepler measure $\nu=C^{\star}(C, R, \psi)$. In particular, for every continuous function $f$ from $O(C, R)$ to $\mathbf{C}$, we have

$$
\begin{equation*}
\frac{1}{n} \sum_{k=0}^{n-1} f\left(S^{k}(z)\right) \xrightarrow[n \rightarrow \infty]{\longrightarrow} \int_{O(C, R)} f \mathrm{~d} \nu \tag{15}
\end{equation*}
$$

uniformly on $O(C, R)$.
In some cases, (15) takes an especially attractive form. Recall that a function $h$ from an open region $\Omega \subseteq \mathbf{R}^{2}$ into $\mathbf{R}$ is harmonic if it is $C^{2}$ and

$$
\frac{\partial^{2} h(x, y)}{\partial x^{2}}+\frac{\partial^{2} h(x, y)}{\partial y^{2}}=0, \quad(x, y) \in \Omega
$$

A function $h: \Omega \rightarrow \mathbf{C}$ is harmonic if both its real part and its imaginary part are harmonic in $\Omega$. (See [30, §11] for more details on harmonic functions.) Denote

$$
D(C, R)=\{z \in \mathbf{C}:|z-C| \leq R\}
$$

and

$$
D^{\circ}(C, R)=\{z \in \mathbf{C}:|z-C|<R\}
$$

Theorem 2.7. In the setup of Proposition 2.6, if $h: D(C, R) \rightarrow \mathbf{C}$ is continuous and is harmonic in $D^{\circ}(C, R)$, then

$$
\frac{1}{N} \sum_{n=0}^{N-1} h\left(\frac{a_{n+1}}{a_{n}}\right) \underset{N \rightarrow \infty}{ } \begin{cases}h\left(\lambda_{1}\right), & \left|c_{2}\right|<\left|c_{1}\right|  \tag{16}\\ h\left(\lambda_{2}\right), & \left|c_{2}\right|>\left|c_{2}\right|\end{cases}
$$

Consider the claim of the theorem for the case where $f$ is the identity function. Suppose, say, that $\left|c_{2}\right|<\left|c_{1}\right|$. The theorem implies

$$
\begin{equation*}
\frac{1}{N} \sum_{n=0}^{N-1} \frac{c_{1} \lambda_{1}^{n+1}+c_{2} \lambda_{2}^{n+1}}{c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}}=\frac{1}{N} \sum_{n=0}^{N-1} \frac{a_{n+1}}{a_{n}} \xrightarrow[N \rightarrow \infty]{ } \lambda_{1} \tag{17}
\end{equation*}
$$

When $\left|c_{2} / c_{1}\right|$ is close to 0 , all terms in the sum on the left-hand side of (17) are very close to $\lambda_{1}$, so that one should expect the sequence to be close to $\lambda_{1}$. The surprising thing is that the limit is exactly $\lambda_{1}$. Moreover, even when $\left|c_{2} / c_{1}\right|$ becomes near (but less than) 1, so that the weight of the term $c_{2} \lambda_{2}^{n}$ is almost as large as that of $c_{1} \lambda_{1}^{n}$, the limit stays $\lambda_{1}$. One may say that, on average, the term $c_{2} \lambda_{2}^{n}$ has no effect.

## 3. Extended Cauchy and circular Cauchy distributions

In this section, we will first recall: (i) the basics of Möbius transformations and their dynamics, and (ii) the Cauchy distribution and its analogue on the

## DANIEL BEREND - RISHI KUMAR

unit circle $S^{1}$. Next, we will discuss at length the distributions defined by the density functions in (8) and (9).

### 3.1. Möbius Transformations

A Möbius transformation $M$ is a mapping of the form

$$
\begin{equation*}
M(z)=\frac{a z+b}{c z+d}, \quad(a, b, c, d \in \mathbf{C}, a d-b c \neq 0) \tag{18}
\end{equation*}
$$

of the extended complex plane $\mathbf{C}_{\infty}=\mathbf{C} \cup\{\infty\}$ into itself. The image of a line or a circle under a Möbius transformation is again a line or circle:

- When $|c|=|d|$, the image of $S^{1}$ under $M$ is the line passing through $(a / c+b / d) / 2$, and making an angle of $\arg (i(a / c-b / d))$ with the positive real axis (see the proof of Theorem 3.1 (i)).
- When $|c| \neq|d|$, the image of $S^{1}$ under $M$ is $O(C, R)$, where $C$ and $R$ are given by

$$
\begin{equation*}
C=\frac{a \bar{c}-b \bar{d}}{|c|^{2}-|d|^{2}}, \quad R=\left|\frac{a d-b c}{|c|^{2}-|d|^{2}}\right|, \tag{19}
\end{equation*}
$$

(see [4, §41-44]). The set of all Möbius transformations is a group under composition. (For more delails on Möbius transformations, we refer to [4].)

### 3.2. The Dynamics of Möbius transformations

Let $M_{1}$ and $M_{2}$ be two Möbius transformations. The transformations are conjugate if $M_{2}=M \circ M_{1} \circ M^{-1}$ for some Möbius transformation $M$. Any Möbius transformation $M \neq I$ has precisely two fixed points in the extended complex plane $\mathbf{C}_{\infty}$, counting multiplicities [3, Theorem 2.6.2]. If $M$ has a single fixed point, say $\zeta$, then $M^{n}(z) \xrightarrow[n \rightarrow \infty]{ } \zeta$ for all $z \in \mathbf{C}_{\infty}$. If $M$ has two distinct fixed points $\zeta_{1}$ and $\zeta_{2}$, then $M$ is conjugate to a Möbius transformation $M_{a}$ of the form

$$
\begin{equation*}
M_{a}(z)=a z, \quad z \in \mathbf{C}_{\infty} \tag{20}
\end{equation*}
$$

for some $a \neq 1$. Therefore, the sequence $\left(M^{n}(z)\right)_{n=0}^{\infty}$ either

1. converges to one of the fixed points of $M$, say $\zeta_{1}$, for all $z \neq \zeta_{2}$ (corresponding to $|a| \neq 1$ in (20)), or
2. moves cyclically through a finite set of points for all $z$ (if $a$ is a root of unity), or
3. forms a dense subset of some line or circle (if $|a|=1$, but $a$ is not a root of unity).
(We refer to 3] for more details.)

### 3.3. Extensions and analogues of the Cauchy distribution

The standard Cauchy distribution is the probability distribution on the real line, defined by the density function

$$
\begin{equation*}
f(x)=\frac{1}{\pi\left(1+x^{2}\right)}, \quad x \in \mathbf{R} \tag{21}
\end{equation*}
$$

Allowing translations of Cauchy random variables, we get the generalized Cauchy distribution, with a density function of the form (see [29])

$$
f(x)=\frac{1}{\pi\left(1+(x-\mu)^{2}\right)}, \quad x \in \mathbf{R},(\mu \in \mathbf{R})
$$

A further generalization is obtained by allowing a scale change as well. For real $\mu, \sigma$, with $\sigma>0$, a random variable $Y$ is Cauchy distributed, with median $\mu$ and scale $\sigma$, if $(Y-\mu) / \sigma$ has a standard Cauchy distribution [32]. McCullagh [26] combined the two parameters into a single complex-valued parameter $\theta=\mu+i \sigma$ in the upper half-plane. The density function is given by

$$
\begin{equation*}
f_{Y}(x)=\frac{\sigma}{\pi|x-\theta|^{2}}, \quad x \in \mathbf{R} . \tag{22}
\end{equation*}
$$

In fact, McCullagh found it convenient to let $\sigma$ be negative also. (The numerator on the right-hand side of (22) is then replaced by $|\sigma|$.) Note that $\theta$ and $\bar{\theta}$ give rise to the same distribution. Also, in the degenerate case $\sigma=0$, the distribution reduces to a point mass at $\mu$.

In Section 2, we have introduced a further generalization, allowing the distribution to be supported on any line $L$ in $\mathbf{C}$. The median $\mu$ is now any complex number, $\sigma$ is still the scale, and we add a third parameter $\alpha$ indicating the direction of $L$ with respect to the positive real axis. (In principle, $\alpha$ ranges over $[0, \pi$ ), but it will be more convenient for us to let it range over $[0,2 \pi)$.) One readily checks that (8) defines the density function of this distribution.

Let $Y$ be a Cauchy distributed random variable, with parameter $\theta=\mu+i \sigma$. Since the transformation

$$
x \rightarrow \frac{i x+1}{-i x+1}, \quad x \in \mathbf{R}
$$

maps $\mathbf{R}$ into $S^{1}$, the complex-valued random variable

$$
Z=\frac{i Y+1}{-i Y+1}
$$

is supported on $S^{1}$. The density of $Z$ is given by

$$
\begin{equation*}
f_{Z}(z)=\frac{\left|1-|\psi|^{2}\right|}{2 \pi|z-\psi|^{2}}, \quad z \in S^{1} \tag{23}
\end{equation*}
$$

where $\psi=(i \theta+1) /(-i \theta+1)($ see [26] $)$.

The distribution defined by this density function is the circular Cauchy distribution, and we write $Z \sim C^{\star}(\psi)$. (The parameter $\psi$ was referred to as eccentricity in Section 2)

The circular Cauchy distribution (a.k.a. the wrapped Cauchy distribution) is a lesser known distribution, although defined already by Lévy [24]. The family of circular Cauchy distributions is closed under Möbius transformations [26]. Moreover, circular Cauchy distributions enjoy the following properties [12, 26]:

1) $C^{\star}(0)$ is the uniform measure on $S^{1 .}$
2) For every $\psi$, the distributions $C^{\star}(\psi)$ and $C^{\star}(1 / \bar{\psi})$ coincide. Thus, it suffices to consider $\psi$-s in the unit disc.
3) As $|\psi|$ increases from 0 to 1 , the distribution deviates more and more from the uniform distribution and becomes concentrated near $\psi /|\psi|$. As $\psi \rightarrow \psi_{0}$ for some $\psi_{0} \in S^{1}$, the distribution converges to a point mass at $\psi_{0}$.
4) If $M$ is the Möbius transformation defined by $M(z)=\beta_{0} z$ with $\beta_{0} \in S^{1}$, then $M\left(C^{\star}(\psi)\right)=C^{\star}\left(\beta_{0} \psi\right)$.
5) If $M$ is the Möbius transformation defined by $M(z)=\left(z+\beta_{1}\right) /\left(\bar{\beta}_{1} z+1\right)$ with $\beta_{1} \in \mathbf{C}$, then $M\left(C^{\star}(\psi)\right)=C^{\star}\left(\left(\psi+\beta_{1}\right) /\left(\overline{\beta_{1}} \psi+1\right)\right)$.
6) If $M(z)$ is the Möbius transformation defined by $M(z)=\beta_{0} \cdot\left(z+\beta_{1}\right) /\left(\bar{\beta}_{1} z+1\right)$ with $\beta_{0} \in S^{1}$ and $\beta_{1} \in \mathbf{C}$, then $M\left(C^{\star}(0)\right)=C^{\star}\left(\beta_{0} \beta_{1}\right)$. (This property follows directly from the two preceding ones, but it will be convenient to have it handy.)
7) If $Z_{1}$ and $Z_{2}$ are independent and $Z_{1} \sim C^{\star}\left(\psi_{1}\right), Z_{2} \sim C^{\star}\left(\psi_{2}\right)$ with $\left|\psi_{1}\right|$, $\left|\psi_{2}\right| \leq 1$, then $Z_{1} Z_{2} \sim C^{\star}\left(\psi_{1} \psi_{2}\right)$.
The following example helps understanding the property 3) better.
Example. In Figure 3, we have "depicted" $C^{\star}(\psi)$ for four values of the parameter $\psi_{l}=0.2 l \cdot \exp (i \pi / 4), 1 \leq l \leq 4$. We have started with the 300 points $\exp (2 \pi i k / 300), 0 \leq k \leq 299$. The (discrete) uniform distribution over these 300 points approximates the uniform distribution over $S^{1}$, which is $C^{\star}(0)$. By the property 5), the measure $C^{\star}(0)$ is taken under the Möbius transformation

$$
M_{\psi}(z)=(z+\psi) /(\bar{\psi} z+1), \quad z \in \mathbf{C}_{\infty}, \quad \text { to } C^{\star}(\psi)
$$

Thus, for each of the four values above of $\psi$, we have drawn the images of those 300 points under $M_{\psi}$. The uniform measure over these points is an approximation of $C^{\star}(\psi)$.

For more information on the circular Cauchy distribution, see [12, 24, 26, 34].


Figure 3. The distribution $C^{\star}(\psi)$ for several values of $\psi$. Note how $|\psi|$ and $\arg (\psi)$ are reflected by the figures.

In Section 2, we have defined a generalization of the circular Cauchy distribution, allowing the distribution to be supported on any circle in $\mathbf{C}$. We added two new parameters; location - the center of the circle, and scale - the radius of the circle. The eccentricity means the same as in the case of an $S^{1}$-supported circular Cauchy distribution. One readily checks that (9) defines the density function of this distribution.

## DANIEL BEREND - RISHI KUMAR

Our main result in this section specifies how Möbius transformations act on the distribution $C^{\star}(0)$. We will explain below that it allows finding how they act on any $C(\mu, \sigma, \alpha)$ and $C^{\star}(C, R, \psi)$.
Theorem 3.1. Let $M$ be any Möbius transformation:

$$
M(z)=\frac{a z+b}{c z+d}, \quad z \in \mathbf{C}_{\infty}, \quad(a, b, c, d \in \mathbf{C}, a d-b c \neq 0)
$$

(i) If $|c|=|d|$, then

$$
M\left(C^{\star}(0)\right)=C(\mu, \sigma, \alpha)
$$

where

$$
\mu=\frac{1}{2}\left(\frac{a}{c}+\frac{b}{d}\right), \quad \sigma=\frac{1}{2}\left|\frac{a}{c}-\frac{b}{d}\right|, \quad \alpha=\arg \left(i\left(\frac{a}{c}-\frac{b}{d}\right)\right) .
$$

(ii) If $|c| \neq|d|$, then

$$
M\left(C^{\star}(0)\right)=C^{\star}(C, R, \psi)
$$

where $C$ and $R$ are as in (19) and

$$
\psi=\frac{\|\left. c\right|^{2}-|d|^{2} \mid}{|c|^{2}-|d|^{2}} \cdot \frac{b c-a d}{|b c-a d|} \cdot \frac{\bar{c}}{d} .
$$

Proof.
(i) In this case $d \neq 0$, so we may rewrite $M$ in the form $M=M_{1} \circ M_{2}$, where

$$
M_{1}(z)=\frac{\frac{a}{c} z+\frac{b}{d}}{z+1}, \quad M_{2}(z)=\frac{c}{d} z, \quad z \in \mathbf{C}_{\infty}
$$

By the property 4) above

$$
\begin{equation*}
M\left(C^{\star}(0)\right)=M_{1} \circ M_{2}\left(C^{\star}(0)\right)=M_{1}\left(C^{\star}(0)\right) \tag{24}
\end{equation*}
$$

The Möbius transformation

$$
T_{2}(z)=\frac{i z+1}{-i z+1}, \quad z \in \mathbf{C}_{\infty}
$$

maps $\mathbf{R} \cup\{\infty\}$ onto $S^{1}$. Therefore, the image of $S^{1}$ under $M_{1}$ is same as the image of $\mathbf{R} \cup\{\infty\}$ under $M_{1} \circ T_{2}$ :

$$
\begin{align*}
M_{1} \circ T_{2}(x) & =\frac{\frac{a}{c}\left(\frac{i x+1}{-i x+1}\right)+\frac{b}{d}}{\frac{i x+1}{-i x++1}+1} \\
& =\frac{1}{2}\left(\frac{a}{c}+\frac{b}{d}\right)+i \frac{1}{2}\left(\frac{a}{c}-\frac{b}{d}\right) x  \tag{25}\\
& =\mu+\sigma_{1} x, \quad x \in \mathbf{R} \cup\{\infty\},
\end{align*}
$$

where $\mu=\frac{1}{2}\left(\frac{a}{c}+\frac{b}{d}\right)$ and $\sigma_{1}=i\left(\frac{a}{c}-\frac{b}{d}\right)$. By (25), the image of $\mathbf{R}$ under $M_{1} \circ T_{2}$ is the line $L$ passing through the point $\mu$ and making an angle of $\alpha=\arg \left(\sigma_{1}\right)$
with the positive real axis. By [26] and [5, p.83], under the transformation $T_{2}$, the image of the standard Cauchy distribution $C(0,1,0)$ is $C^{\star}(0)$. Hence we may rewrite (24) in the form

$$
\begin{equation*}
M\left(C^{\star}(0)\right)=M_{1} \circ T_{2}(C(0,1,0)) \tag{26}
\end{equation*}
$$

By (21) and (25), the density function $f_{1}$ of the distribution $M_{1} \circ T_{2}(C(0,1,0))$ on the line $L$ is

$$
\begin{align*}
f_{1}(z) & =f\left(\left(M_{1} \circ T_{2}\right)^{-1}(z)\right)\left|\frac{\mathrm{d}}{\mathrm{~d} z}\left(M_{1} \circ T_{2}\right)^{-1}(z)\right| \\
& =\frac{1}{\pi} \cdot \frac{1}{1+\left|(z-\mu) / \sigma_{1}\right|^{2}}\left|\frac{\mathrm{~d}}{\mathrm{~d} z}\left(\frac{z-\mu}{\sigma_{1}}\right)\right|  \tag{27}\\
& =\frac{1}{\pi\left|\sigma_{1}\right|} \cdot \frac{1}{1+\left|(z-\mu) / \sigma_{1}\right|^{2}}, \quad z \in L .
\end{align*}
$$

Hence, by (8), (26), and (27), we have $M\left(C^{\star}(0)\right)=C(\mu, \sigma, \alpha)$, where $\sigma=\left|\sigma_{1}\right|$.
(ii) Define a Möbius transformation $M_{3}$ by

$$
\begin{equation*}
M_{3}(z)=\frac{z-C}{R}, \quad z \in \mathbf{C}_{\infty} \tag{28}
\end{equation*}
$$

By (19), the Möbius transformation $M_{3} \circ M$ maps $S^{1}$ onto itself. By (19) and (28), we have for $z \in \mathbf{C}$,

$$
\begin{align*}
M_{3} \circ M(z) & =\left(\frac{a z+b}{c z+d}-\frac{a \bar{c}-b \bar{d}}{|c|^{2}-|d|^{2}}\right) \cdot \frac{\left.| | c\right|^{2}-|d|^{2} \mid}{|b c-a d|} \\
& =\frac{\|\left. c\right|^{2}-|d|^{2} \mid}{|c|^{2}-|d|^{2}} \cdot \frac{1}{|b c-a d|} \cdot \frac{(a z+b)\left(|c|^{2}-|d|^{2}\right)-(c z+d)(a \bar{c}-b \bar{d})}{c z+d} \\
& =\frac{|c|^{2}-|d|^{2} \mid}{|c|^{2}-|d|^{2}} \cdot \frac{b c-a d}{|b c-a d|} \cdot \frac{\bar{d} z+\bar{c}}{c z+d} \\
& =\frac{|c|^{2}-|d|^{2} \mid}{|c|^{2}-|d|^{2}} \cdot \frac{b c-a d}{|b c-a d|} \cdot \frac{\bar{d}}{d} \cdot \frac{z+\bar{c} / \bar{d}}{(c / d) z+1} . \tag{29}
\end{align*}
$$

By the property 6) above and (29),

$$
\begin{equation*}
M_{3} \circ M\left(C^{\star}(0)\right)=C^{\star}(\psi) \tag{30}
\end{equation*}
$$

where $\psi=\frac{\left||c|^{2}-|d|^{2}\right|}{|c|^{2}-|d|^{2}} \cdot \frac{b c-a d}{|b c-a d|} \cdot \frac{\bar{c}}{d}$. By (30),

$$
\begin{equation*}
M\left(C^{\star}(0)\right)=M_{3}^{-1} C^{\star}(\psi) \tag{31}
\end{equation*}
$$

The density function $f_{2}$ of the distribution $C^{\star}(\psi)$ may be written in the form

$$
\begin{equation*}
f_{2}(\exp (i \theta))=\frac{\left|1-|\psi|^{2}\right|}{2 \pi|\exp (i \theta)-\psi|^{2}}, \quad \theta \in[0,2 \pi) \tag{32}
\end{equation*}
$$

By (32), the density function $f_{3}$ of the distribution $M_{3}^{-1}\left(C^{\star}(\psi)\right)$ on $O(C, R)$ is

$$
\begin{align*}
f_{3}(C+R \exp (i \theta)) & =f_{2}\left(M_{3}(C+R \exp (i \theta))\right)\left|\frac{1}{R} \cdot \frac{\mathrm{~d}}{\mathrm{~d} \theta}\left(M_{3}(C+R \exp (i \theta))\right)\right| \\
& =\frac{\left|1-|\psi|^{2}\right|}{2 \pi|\exp (i \theta)-\psi|^{2}} \cdot\left|\frac{1}{R} \cdot \frac{\mathrm{~d}}{\mathrm{~d} \theta}(\exp (i \theta))\right| \\
& =\frac{\left|1-|\psi|^{2}\right|}{2 \pi|\exp (i \theta)-\psi|^{2}} \cdot \frac{1}{R}, \quad \theta \in[0,2 \pi) \tag{33}
\end{align*}
$$

Hence, by (9), (31), and (33), we have

$$
M\left(C^{\star}(0)\right)=M_{3}^{-1}\left(C^{\star}(\psi)\right)=C^{\star}(C, R, \psi)
$$

Note that we have stated in Theorem 3.1 only to what measure the uniform distribution $C^{\star}(0)$ on $S^{1}$ is mapped under any Möbius transformation. This allows us finding the image of any Cauchy measure and circular Cauchy measure. Indeed, let $L$ be a line in $\mathbf{C}$, endowed with measure $C(\mu, \sigma, \alpha)$, and $M$ any Möbius transformation. By Theorem 3.1. (i), letting $M_{\mu, \sigma, \alpha}(z)=(a z+b) /(z+1)$, where $a=\mu+\exp (i(\alpha-\pi / 2)) \sigma$ and $b=\mu-\exp (i(\alpha-\pi / 2)) \sigma$, we have

$$
M_{\mu, \sigma, \alpha}\left(C^{\star}(0)\right)=C(\mu, \sigma, \alpha)
$$

Therefore, $M(C(\mu, \sigma, \alpha))=M \circ M_{\mu, \sigma, \alpha}\left(C^{\star}(0)\right)$. Similarly, let $O(C, R)$ be any circle, endowed with the measure $C^{\star}(C, R, \psi)$. It follows from the property 6) and Theorem 3.1. (ii) that

$$
M\left(C^{\star}(C, R, \psi)\right)=M \circ M_{3}^{-1} M_{\psi}\left(C^{\star}(0)\right)
$$

where $M_{3}(z)=(z-C) / R$ and $M_{\psi}(z)=(z+\psi) /(\bar{\psi} z+1)$.
We mention in passing that the circular Cauchy (as well as its generalization to higher-dimensional spheres [5, 6, 9, 11]) distribution is usually used to model angular data. Examples of such data are migration of turtles [28], orientation of ants towards a black target [1,28, and wind direction data [1, 12]. The data lies in $[0,2 \pi]$, but it makes little sense to treat it as real data in this interval. It is more natural to view it as data on the unit circle in the plane. Thus, viewing the data as corresponding to complex numbers on the unit circle is a matter of convenience, but the data is not really complex-valued. In our case, though, the circular Cauchy distribution really describes the behaviour of complex-valued data.

## 4. Discussion and Intuition

Suppose we fix $\lambda_{1}$ and $\lambda_{2}$, with $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$ and $\lambda_{2} / \lambda_{1}$ not a root of unity. What can be said about the Kepler sets of all sequences $a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}$, as $c_{1}, c_{2}$ vary over $\mathbf{C}$ ? By Remark [1, only the absolute ratio $r=\left|c_{2} / c_{1}\right|$ plays a role in determining the Kepler set. Without loss of generality, we may therefore restrict our attention to $\left(a_{n}\right)$ of the form

$$
a_{n}=\lambda_{1}^{n}+r \lambda_{2}^{n}, \quad n \geq 0, \quad(r \in(0, \infty)) .
$$

Let us describe how the Kepler set changes as $r$ varies from 0 to $\infty$. For $r \approx 0$, by (7), the circle $O(C, R)$ has its center close to $\lambda_{1}$ and its radius close to 0 . Indeed,

$$
\frac{a_{n+1}}{a_{n}}=\frac{\lambda_{1}^{n+1}+r \lambda_{2}^{n+1}}{\lambda_{1}^{n}+r \lambda_{2}^{n}} \approx \frac{\lambda_{1}^{n+1}}{\lambda_{1}^{n}}=\lambda_{1},
$$

which shows that the consecutive ratios become close to $\lambda_{1}$.
Rewrite the formula for $C$ in (7) in the form:

$$
\begin{equation*}
\lambda_{1}=\left(1-r^{2}\right) C+r^{2} \lambda_{2} \tag{34}
\end{equation*}
$$

Thus, for $0<r<1$, it follows from (34) that $\lambda_{1}$ is a convex combination of $C$ and $\lambda_{2}$. Put differently, $C$ lies on the line passing through $\lambda_{1}$ and $\lambda_{2}$, so that $\lambda_{1}$ is between $\lambda_{2}$ and $C$. It will follow from the proof of Theorem 2.2 that the Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is the image of $S^{1}$ under the Möbius transformation $T(z)=\frac{r \lambda_{2} z+\lambda_{1}}{r z+1}$. The Möbius transformation defined by

$$
M_{3}(z)=\frac{1}{R}(z-C), \quad z \in \mathbf{C}_{\infty}
$$

maps $O(C, R)$ to $S^{1}$. By (19),

$$
\begin{aligned}
M_{3}(T(z)) & =\frac{\left|1-r^{2}\right|}{r\left|\lambda_{2}-\lambda_{1}\right|} \cdot\left(\frac{\lambda_{1}+r \lambda_{2} z}{1+r z}-\frac{\lambda_{1}-r^{2} \lambda_{2}}{1-r^{2}}\right) \\
& =\frac{\left|1-r^{2}\right|}{1-r^{2}} \cdot \frac{\lambda_{2}-\lambda_{1}}{\left|\lambda_{2}-\lambda_{1}\right|} \cdot \frac{z+r}{r z+1} \\
& =\alpha \cdot \frac{z+r}{r z+1}, \quad z \in S^{1}
\end{aligned}
$$

where $\alpha=\frac{\left|1-r^{2}\right|}{1-r^{2}} \cdot \frac{\lambda_{2}-\lambda_{1}}{\left|\lambda_{2}-\lambda_{1}\right|} \in S^{1}$. Since $M_{3}$ maps $D^{\circ}(C, R)$ onto $D^{\circ}(0,1)$, and the complement of $D^{\circ}(C, R)$ onto the complement of $D^{\circ}(0,1)$, we have

$$
M_{3}(T(0))=\alpha r \quad \text { and } \quad\left|M_{3}(T(0))\right|=r<1
$$

Hence, $T(0)=\lambda_{1}$ is inside $O(C, R)$ and $T(\infty)=\lambda_{2}$ is outside.


Figure 4. The Kepler sets of the sequences $a_{n}=\lambda_{1}^{n}+r \lambda_{2}^{n}$, where $\lambda_{1}=(3+4 i) / 5, \lambda_{2}=(5-12 i) / 13$, with various values of $r$. For $r=1$, the Kepler set is the line $L$. For $r=0.2,0.4,0.6,0.8$ we get four circles above $L$, and for $r=1 / 0.2,1 / 0.4,1 / 0.6,1 / 0.8-$ their mirror images with respect to $L$.

As $r$ increases from 0 to 1 , the center gets further away from $\lambda_{1}$ and the radius increases. As $r \rightarrow 1^{-}$, both the center and the radius tend to $\infty$. By Theorem 2.4, in the process, $|\psi|$ increases from 0 to 1 , while $\arg (\psi)=\arg \left(\lambda_{2}-\lambda_{1}\right)$ remains fixed. Thus, the distribution we get on $O(C, R)$ changes gradually from a uniform measure (in the limit as $r \rightarrow 0$ ) to become more and more concentrated near the intersection of $O(C, R)$ with the interval $\left[\lambda_{1}, \lambda_{2}\right]$. The density on $O(C, R)$ is largest at the point closest to $\lambda_{2}$, and becomes smaller as we get further from $\lambda_{2}$. (See Figure 4)

The circles we obtain, considered as circles in the Riemann sphere, are pairwise disjoint and converge to a great circle as $r \rightarrow 1^{-}$. This great circle is the Kepler set for $r=1$, which is the line mean perpendicular to the interval $\left[\lambda_{1}, \lambda_{2}\right]$ (see Figure 4).

For $r>1$, the situation is similar. The Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is the same as that of $\frac{1}{r} a_{n}=\lambda_{2}^{n}+\frac{1}{r} \lambda_{1}^{n}$, so that we get the same picture as for $r<1$, where $\lambda_{1}, \lambda_{2}$ switch roles.

Remark 3. Since the Kepler sets are invariant under the transformation $S$, the transformation $S$ partitions $\mathbf{C}_{\infty}$ into disjoint $S$-invariant circles and an $S$ --invariant line. Thus, $S$ is a semi-simple transformation according to the terminology in [33, p. 121].

We explain intuitively why only $\left|c_{2} / c_{1}\right|$ is important. Shifting the sequence, we may consider the sequence $a_{n+k}=c_{1} \lambda_{1}^{k} \lambda_{1}^{n}+c_{2} \lambda_{2}^{k} \lambda_{2}^{n}$ with arbitrary $k$ instead of $a_{n}$. Thus, the coefficients $c_{1}, c_{2}$ give rise to the same Kepler set as do the coefficients $c_{1} \lambda_{1}^{k}, c_{2} \lambda_{2}^{k}$ for any $k$. Taking a sequence $\left(k_{i}\right)_{i=0}^{\infty}$ for which $\left(c_{2} / c_{1}\right)\left(\lambda_{2} / \lambda_{1}\right)^{k_{i}} \xrightarrow[i \rightarrow \infty]{ }\left|c_{2} / c_{1}\right|$, we see that $c_{1}, c_{2}$ may be replaced by $\left|c_{1}\right|,\left|c_{2}\right|$.

## 5. Proofs

Proof of Lemma 2.1. The characteristic polynomial of $\left(a_{n}\right)_{n=0}^{\infty}$ is given by $f(x)=x^{2}-c x-d$, and its roots are $\lambda_{1,2}=\left(c \pm \sqrt{c^{2}+4 d}\right) / 2$. If $c=0$ or $c^{2}+4 d=0$, then $\lambda_{2} / \lambda_{1}$ is a root of unity. Also, $\lambda_{1}$ and $\lambda_{2}$ cannot be both real, as otherwise the equality $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$ would imply that $c=0$.

Since $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$, the line segment $\left[\lambda_{1}, \lambda_{2}\right]$ is perpendicular to $[0, c / 2]$. It follows that $\sqrt{c^{2}+4 d} / c \in i \mathbf{R}$, so that $\left(c^{2}+4 d\right) / c^{2} \in\{x: x<0\}$. Hence $d<-c^{2} / 4$.

Write $a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}$ for suitable non-zero $c_{1}, c_{2}$, depending on the initial values $a_{0}, a_{1}$. Suppose that $a_{n}=0$ for some $n$. Then

$$
\begin{equation*}
\left(\lambda_{2} / \lambda_{1}\right)^{n}=\left(-c_{1} / c_{2}\right) \tag{35}
\end{equation*}
$$

If $\lambda_{2} / \lambda_{1}$ is not a root of unity, then (35) can have at most one solution $n \in \mathbf{Z}$.
Proof of Theorem 2.2, 1. We have

$$
\begin{equation*}
\frac{a_{n+1}}{a_{n}}=\lambda_{1} \cdot \frac{c_{2} / c_{1} \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n+1}+1}{c_{2} / c_{1} \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n}+1} . \tag{36}
\end{equation*}
$$

Since $\lambda_{2} / \lambda_{1}$ is not a root of unity, the sequence $\left(\left(\lambda_{2} / \lambda_{1}\right)^{n}\right)_{n=0}^{\infty}$ is dense in $S^{1}$ (see [31, Proposition 1.3.4]). Therefore, the Kepler set of $\left(a_{n}\right)_{n=0}^{\infty}$ is

$$
\begin{equation*}
K=\left\{\lambda_{1} \cdot \frac{\left(c_{2} / c_{1}\right)\left(\lambda_{2} / \lambda_{1}\right) z+1}{\left(c_{2} / c_{1}\right) z+1}: z \in S^{1}\right\} \tag{37}
\end{equation*}
$$

We now continue separately in the two cases $\left|c_{1}\right|=\left|c_{2}\right|$ and $\left|c_{1}\right| \neq\left|c_{2}\right|$.
If $\left|c_{1}\right|=\left|c_{2}\right|$, as $z$ varies over $S^{1}$, so does $\left(c_{2} / c_{1}\right) z$. Hence we may write $K$ in a simpler form

$$
\begin{equation*}
K=\left\{\frac{\lambda_{2} z+\lambda_{1}}{z+1}: z \in S^{1}\right\} \tag{38}
\end{equation*}
$$

It follows that $K=T_{1}\left(S^{1}\right)$, where $T_{1}$ is the Möbius transformation given by

$$
\begin{equation*}
T_{1}(z)=\frac{\lambda_{2} z+\lambda_{1}}{z+1}, \quad z \in \mathbf{C}_{\infty} \tag{39}
\end{equation*}
$$

## DANIEL BEREND - RISHI KUMAR

Let $T_{2}$ be the Möbius transformation defined by

$$
\begin{equation*}
T_{2}(z)=\frac{i z+1}{-i z+1}, \quad z \in \mathbf{C}_{\infty} \tag{40}
\end{equation*}
$$

and notice that it takes $\mathbf{R} \cup\{\infty\}$ onto $S^{1}$. Thus, $K=T_{1} \circ T_{2}(\mathbf{R} \cup\{\infty\})$. Now

$$
\begin{equation*}
T_{1} \circ T_{2}(x)=\frac{\lambda_{1}+\lambda_{2}}{2}\left(i \cdot \frac{\lambda_{2}-\lambda_{1}}{\lambda_{1}+\lambda_{2}} x+1\right), \quad x \in \mathbf{R} \cup\{\infty\} \tag{41}
\end{equation*}
$$

Since $i \cdot \frac{\lambda_{2}-\lambda_{1}}{\lambda_{1}+\lambda_{2}}$ is real, the image of $\mathbf{R} \cup\{\infty\}$ under $T_{1} \circ T_{2}$ is the line passing through the origin and the point $\left(\lambda_{1}+\lambda_{2}\right) / 2$.

Now consider the case $\left|c_{1}\right| \neq\left|c_{2}\right|$. Let $c_{2} / c_{1}=r \exp (i \phi)$ for some $r>0$ and $\phi \in[0,2 \pi)$, with $r \neq 1$. Replacing $z$ in (37) by $\exp (i \phi) z$, we see that

$$
K=\left\{\frac{r \lambda_{2} z+\lambda_{1}}{z+1}: z \in S^{1}\right\}
$$

Denoting by $T$ the Möbius transformation given by

$$
\begin{equation*}
T(z)=\frac{r \lambda_{2} z+\lambda_{1}}{r z+1}, \quad z \in \mathbf{C}_{\infty} \tag{42}
\end{equation*}
$$

we obtain $K=T\left(S^{1}\right)$. By [4, §41-44] (see the result mentioned in Subsection 3.1), $T\left(S^{1}\right)=O(C, R)$, where the center $C$ and radius $R$ are given by

$$
\begin{equation*}
C=\frac{\lambda_{1}-r^{2} \lambda_{2}}{1-r^{2}}=\frac{\left|c_{1}\right|^{2} \lambda_{1}-\left|c_{2}\right|^{2} \lambda_{2}}{\left|c_{1}\right|^{2}-\left|c_{2}\right|^{2}} \tag{43}
\end{equation*}
$$

and

$$
\begin{equation*}
R=\frac{\left|r \lambda_{1}-r \lambda_{2}\right|}{\left|1-r^{2}\right|}=\frac{\left|c_{1}\right|\left|c_{2}\right|\left|\lambda_{1}-\lambda_{2}\right|}{\left|\left|c_{1}\right|^{2}-\left|c_{2}\right|^{2}\right|} . \tag{44}
\end{equation*}
$$

A routine calculation shows now that

$$
\begin{equation*}
R^{2}+\left|\lambda_{1}\right|^{2}=|C|^{2} \tag{45}
\end{equation*}
$$

Thus, $R<|C|$, so that that the origin is outside $O(C, R)$.
2. Let $L$ be a line passing through the origin and making an angle $\theta$ with the positive real axis. Choose $\lambda_{1}=\exp (i(\theta-\varepsilon))$ and $\lambda_{2}=\exp (i(\theta+\varepsilon))$ for some $\varepsilon \in(0, \pi / 2)$ which is not a rational multiple of $\pi$. By part 1. (i), the line $L$ is the Kepler set of the sequence $\left(a_{n}\right)_{n=0}^{\infty}$, given by $a_{n}=\lambda_{1}^{n}+\lambda_{2}^{n}$ for each $n$.

Let $O(C, R)$ be a circle such that $R<|C|$. We need to construct a second--order linear recurrence sequence $\left(a_{n}\right)_{n=0}^{\infty}$ whose Kepler set is $O(C, R)$. It is easily seen that the circles $O\left(0, \sqrt{|C|^{2}-R^{2}}\right)$ and $O(C, R)$ intersect orthogonally. (Refer to Figure 5 in what follows.) Draw through $C$ a line intersecting $O\left(0, \sqrt{|C|^{2}-R^{2}}\right)$ at two points. Let $\lambda_{1}$ be the intersection point closer to $C$, and $\lambda_{2}$ - the one farther away. Perturbing this line if necessary, we may assume that $\lambda_{2} / \lambda_{1}$ is not a root of unity. We can write $\lambda_{1}$ as a convex combination of $\lambda_{2}$ and $C$,

## CONSECUTIVE RATIOS IN RECURRENCE SEQUENCES



Figure 5. Construction of a second-order recurrence sequence whose Kepler set is the dashed red circle $O(2+2 i, 3 / 2)$. The roots need to lie on the dotted black circle $O\left(0, \sqrt{|2+2 i|^{2}-(3 / 2)^{2}}\right)=O(0, \sqrt{23} / 2)$.
say $\lambda_{1}=\left(1-r^{2}\right) C+r^{2} \lambda_{2}$ for some $r \in(0,1)$. Since $\left|\lambda_{1}\right|^{2}=\left|\lambda_{2}\right|^{2}=|C|^{2}-R^{2}$, we have $R=r \frac{\left|\lambda_{1}-\lambda_{2}\right|}{1-r^{2}}$. It follows from Theorem [2.2, 1. (ii) and (42), (43), (44) that $O(C, R)$ is the Kepler set of the sequence $\left(a_{n}\right)_{n=0}^{\infty}$, given by $a_{n}=\lambda_{1}^{n}+r \lambda_{2}^{n}$ for each $n$.

In the next proof, we will make use of
Proposition 5.1 ([23, p.178]). Let $\left(x_{n}\right)_{n=0}^{\infty}$ be a sequence in a compact Hausdorff space $X$, distributed according to some probability measure $\nu$ on $X$. Let $Y$ be another compact Hausdorff space and $f: X \rightarrow Y$ a continuous function. Then the sequence $\left(f\left(x_{n}\right)\right)_{n=0}^{\infty}$ is distributed in $Y$ according to $f(\nu)$.

The proof is immediate.
Proof of Theorem [2.4, 1. Refer again to (36):

$$
\begin{equation*}
\frac{a_{n+1}}{a_{n}}=\lambda_{1} \cdot \frac{c_{2} / c_{1} \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n+1}+1}{c_{2} / c_{1} \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n}+1} \tag{36}
\end{equation*}
$$

The sequence $\left(c_{2} / c_{1} \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n}\right)_{n=0}^{\infty}$ is distributed according to the uniform measure $C^{\star}(0)$ on $S^{1}$ (see [23, Example 2.1]). By Proposition 5.1, the Kepler measure of the sequence $\left(a_{n}\right)_{n=0}^{\infty}$ is $T_{1}\left(C^{\star}(0)\right)$, where $T_{1}(z)=\frac{\lambda_{2} z+\lambda_{1}}{z+1}$. From Theorem 3.1 we obtain $T_{1}\left(C^{\star}(0)\right)=C(\mu, \sigma, \alpha)$, with $\mu=\frac{\lambda_{1}+\lambda_{2}}{2}, \sigma=\left|\frac{\lambda_{2}-\lambda_{1}}{2}\right|$, and $\alpha=\arg \left(i\left(\lambda_{1}-\lambda_{2}\right)\right)=\arg \left(\lambda_{1}+\lambda_{1}\right)=\arg (\mu)$.
2. Rewrite (36) in the form

$$
\frac{a_{n+1}}{a_{n}}=\frac{r \cdot \exp (i \theta) \cdot \lambda_{2}\left(\lambda_{2} / \lambda_{1}\right)^{n}+\lambda_{1}}{r \cdot \exp (i \theta) \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n}+1}
$$

where $c_{2} / c_{1}=r \exp (i \theta)$ for some $r>0$ and $\theta \in(0,2 \pi)$. Thus the sequence $\left(\exp (i \theta) \cdot\left(\lambda_{2} / \lambda_{1}\right)^{n}\right)_{n=0}^{\infty}$ is distributed according to $C^{\star}(0)$ on $S^{1}$. By Proposition 5.1. Theorem 3.1, (42), (431), and (44), the Kepler measure of the sequence $\left(a_{n}\right)_{n=0}^{\infty}$ is $T\left(C^{\star}(0)\right)=C^{\star}(C, R, \psi)$, where

$$
T(z)=\frac{r \lambda_{2} z+\lambda_{1}}{r z+1}, \quad \text { and } \quad \psi=r \cdot \frac{\left|1-r^{2}\right|}{1-r^{2}} \cdot \frac{\lambda_{2}-\lambda_{1}}{\left|\lambda_{2}-\lambda_{1}\right|}
$$

Proof of Theorem [2.5, 1. Only if: Let $\left(a_{n}\right)_{n=0}^{\infty}$ be a second-order linear recurrence sequence with Kepler measure $C(\mu, \sigma, \alpha)$. Let $\lambda_{1}$ and $\lambda_{2}$ be the roots of the characteristic polynomial of $\left(a_{n}\right)_{n=0}^{\infty}$. It follows from the discussion in Sections 1 and 2 that $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$ and $\lambda_{2} / \lambda_{1}$ is not a root of unity. By Theorem 2.2. 1. (i), we have $\mu=\left(\lambda_{1}+\lambda_{2}\right) / 2, \sigma=\left|\left(\lambda_{1}-\lambda_{2}\right) / 2\right|$ and $\alpha=\arg (\mu)$. The triangle $\triangle O \lambda_{1} \lambda_{2}$ (where $O$ denote the origin) is an isosceles triangle. The line passing through $O$ and $\mu$ bisects the side $\left[\lambda_{1}, \lambda_{2}\right.$ ] of the triangle. Hence, this line is also a height of the triangle. It follows that $\lambda_{1}=\mu+i(\mu /|\mu|) \sigma$ and $\lambda_{2}=\mu-i(\mu /|\mu|) \sigma$. We have

$$
\frac{\lambda_{2}}{\lambda_{1}}=\frac{1-i \sigma /|\mu|}{1+i \sigma /|\mu|}=\exp (-2 i \arctan (\sigma /|\mu|))
$$

Since $\lambda_{2} / \lambda_{1}$ is not a root of unity, $\arctan (\sigma /|\mu|)$ is not a rational multiple of $\pi$. If: Choose $\lambda_{1}=\mu+i(\mu /|\mu|) \sigma$ and $\lambda_{2}=\mu-i(\mu /|\mu|) \sigma$. We have

$$
\frac{1}{\pi} \arg \left(\lambda_{2} / \lambda_{1}\right)=\frac{-2}{\pi} \arctan (\sigma /|\mu|) \notin \mathbf{Q} .
$$

It follows from Proof of Theorem [2.4. 1. that the Kepler measure of the sequence $\left(a_{n}\right)_{n=0}^{\infty}$, given by $a_{n}=\lambda_{1}^{n}+\lambda_{2}^{n}$ for $n \geq 0$, is $C(\mu, \sigma, \alpha)$.
2. Since $C^{\star}(C, R, \psi)=C^{\star}(C, R, 1 / \bar{\psi})$ for all $\psi \in \mathbf{C}$ with $|\psi| \neq 1$, it suffices to prove the theorem for $|\psi|<1$.
Only if: Let $\left(a_{n}\right)_{n=0}^{\infty}$ be a second-order linear recurrence sequence with Kepler measure $C^{\star}(C, R, \psi)$. Let $\lambda_{1}$ and $\lambda_{2}$ be the roots of the characteristic polynomial of $\left(a_{n}\right)_{n=0}^{\infty}$. As in the first part of the proof, $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$ and $\lambda_{2} / \lambda_{1}$ is not a root of unity. Write $a_{n}$ in the form

$$
a_{n}=c_{1} \lambda_{1}^{n}+c_{2} \lambda_{2}^{n}, \quad n=0,1,2, \ldots
$$

for suitable non-zero $c_{1}$ and $c_{2}$ with $\left|c_{2}\right| \neq\left|c_{1}\right|$. Suppose, say, that, $\left|c_{2}\right|<\left|c_{1}\right|$. Put $r=\left|c_{2} / c_{1}\right|<1$. By Theorem 2.4. 2,

$$
\begin{equation*}
\psi=r \frac{\lambda_{2}-\lambda_{1}}{\left|\lambda_{2}-\lambda_{1}\right|} \tag{46}
\end{equation*}
$$

By (7) and (46),

$$
\begin{align*}
\left|\frac{\psi-C^{\prime}}{R^{\prime}}\right| & =\frac{1}{R^{\prime}}\left|\frac{r\left(\lambda_{2}-\lambda_{1}\right)}{\left|\lambda_{2}-\lambda_{1}\right|}+\frac{C}{R}\right| \\
& =\frac{1}{R R^{\prime}}\left|\frac{r R\left(\lambda_{2}-\lambda_{1}\right)}{\left|\lambda_{2}-\lambda_{1}\right|}+C\right| \\
& =\frac{1}{\left|\lambda_{1}\right|}\left|\frac{r^{2}}{1-r^{2}}\left(\lambda_{2}-\lambda_{1}\right)+\frac{\lambda_{1}-r^{2} \lambda_{2}}{1-r^{2}}\right| \\
& =\frac{1}{\mid \lambda_{1}}\left|\frac{\left(1-r^{2}\right) \lambda_{1}}{\left(1-r^{2}\right)}\right|=1 . \tag{47}
\end{align*}
$$

Hence $\psi$ lies on the circle $O\left(C^{\prime}, R^{\prime}\right)$. By (7), Theorem [2.42, and (46), we have $C+R \psi=\lambda_{1}$ and $C+R / \bar{\psi}=\lambda_{2}$. Hence $\frac{1}{\pi} \arg \left(\lambda_{2} / \lambda_{1}\right)=\frac{1}{\pi} \arg \left(\frac{C+R / \bar{\psi}}{C+R \psi}\right) \notin \mathbf{Q}$. If: Let $\lambda_{1}=C+R \psi$ and $\lambda_{2}=C+R / \bar{\psi}$. Since $\psi$ lies on the circle $O\left(C^{\prime}, R^{\prime}\right)$, we have

$$
\begin{align*}
C \bar{\psi}+\bar{C} \psi & =R\left(-C^{\prime} \bar{\psi}-\bar{C}^{\prime} \psi\right) \\
& =R\left(\left|C^{\prime}-\psi\right|^{2}-\left|C^{\prime}\right|^{2}-|\psi|^{2}\right) \\
& =R\left(R^{\prime 2}-\left|C^{\prime}\right|^{2}-|\psi|^{2}\right) \\
& =R\left(-1-|\psi|^{2}\right) \tag{48}
\end{align*}
$$

By (48),

$$
\begin{align*}
\left|\frac{\lambda_{2}}{\lambda_{1}}\right|^{2} & =\frac{|C+R / \bar{\psi}|^{2}}{|C+R \psi|^{2}} \\
& =\frac{|C|^{2}+R^{2} /|\psi|^{2}+R(C \bar{\psi}+\bar{C} \psi) /|\psi|^{2}}{|C|^{2}-R^{2}} \\
& =\frac{|C|^{2}+R^{2} /|\psi|^{2}+R\left(-R-R|\psi|^{2}\right) /|\psi|^{2}}{|C|^{2}-R^{2}} \\
& =1 \tag{49}
\end{align*}
$$

Hence $\left|\lambda_{1}\right|=\left|\lambda_{2}\right|$, and from (11) it follows that $\lambda_{2} / \lambda_{1}$ is not a root of unity. Define a linear recurrence sequence $\left(a_{n}\right)_{n=0}^{\infty}$ by

$$
a_{n}=\lambda_{1}^{n}+|\psi| \lambda_{2}^{n}, \quad n=0,1,2, \ldots
$$

By Theorem 2.4. 2. the Kepler measure of $\left(a_{n}\right)_{n=0}^{\infty}$ is $C^{\star}(C, R, \psi)$.
Proof of Proposition 2.6. Define the Möbius transformations $R$ and $T$ by

$$
R(z)=\frac{\lambda_{2}}{\lambda_{1}} z, \quad T(z)=\frac{r \lambda_{2} z+\lambda_{1}}{z+1}, \quad \text { for } z \in \mathbf{C}_{\infty}
$$

## DANIEL BEREND - RISHI KUMAR

The uniform measure on $S^{1}$ is the unique $R$-invariant probability measure [31, Proposition 6.4.4]. By (5) and (36), the diagram

is commutative. Therefore, $\nu=C^{\star}(C, R, \psi)$ is the unique $S$-invariant measure [13, p.145], where $\psi$ is as in Theorem [2.4. By (14), for every continuous function $f: O(C, R) \rightarrow \mathbf{C}$,

$$
\begin{equation*}
\frac{1}{N} \sum_{n=0}^{N-1} f\left(S^{n}(x)\right) \underset{n \rightarrow \infty}{\longrightarrow} \int_{O(C, R)} f \mathrm{~d} \nu \tag{51}
\end{equation*}
$$

uniformly on $O(C, R)$.
Proof of Theorem [2.7] It is enough to prove the theorem for real-valued $h$. Let $f_{3}$ be as in the proof of Theorem (3.1. By (33) and (51), for $z \in O(C, R)$

$$
\begin{align*}
\frac{1}{N} \sum_{n=0}^{N-1} h\left(S^{n}(z)\right) \xrightarrow[n \rightarrow \infty]{\longrightarrow} & \int_{O(C, R)} h \mathrm{~d} f_{3} \\
& =\int_{0}^{2 \pi} h(C+R \exp (i \theta)) f_{3}(C+R \exp (i \theta)) R \cdot \mathrm{~d} \theta \\
& =\int_{0}^{2 \pi} h(C+R \exp (i \theta)) \frac{\left|1-|\psi|^{2}\right|}{2 \pi|\exp (i \theta)-\psi|^{2}} \frac{1}{R} \cdot R \mathrm{~d} \theta \\
& =\int_{0}^{2 \pi} h(C+R \exp (i \theta)) \frac{\left|1-|\psi|^{2}\right|}{2 \pi|\exp (i \theta)-\psi|^{2}} \mathrm{~d} \theta \tag{52}
\end{align*}
$$

By [30, Theorem 11.9] and (52), for every $z \in O(C, R)$

$$
\frac{1}{N} \sum_{n=0}^{N-1} h\left(S^{n}(z)\right) \underset{n \rightarrow \infty}{\longrightarrow} h(C+R \psi)= \begin{cases}h\left(\lambda_{1}\right), & \left|c_{2}\right|<\left|c_{1}\right|,  \tag{53}\\ h\left(\lambda_{2}\right), & \left|c_{1}\right|<\left|c_{2}\right|\end{cases}
$$

Putting $z=a_{1} / a_{0}$ in (53), we obtain

$$
\frac{1}{N} \sum_{n=0}^{N-1} h\left(\frac{a_{n+1}}{a_{n}}\right) \underset{n \rightarrow \infty}{\longrightarrow} \begin{cases}h\left(\lambda_{1}\right), & \left|c_{2}\right|<\left|c_{1}\right|  \tag{54}\\ h\left(\lambda_{2}\right), & \left|c_{1}\right|<\left|c_{2}\right|\end{cases}
$$

## CONSECUTIVE RATIOS IN RECURRENCE SEQUENCES

Acknowledgement. The authors express their gratitude to P. McCullagh for lengthy correspondence regarding Cauchy distributions and to J.-P. Allouche for his comments on the first draft of the paper. We also thank the referee for his many helpful comments.

## REFERENCES

[1] ABUZAID, A. H.-EL-HANJOURI, M. M.-KULAB, M. M.: On discordance tests for the wrapped Cauchy distribution, Open J. Statist. 5 (2015), no. 4, 245-253.
[2] BAGDASAR, O.-HEDDERWICK, E.-POPA, I. L.: On the ratios and geometric boundaries of complex Horadam sequences, Electron. Notes Discrete Math. 67 (2018), 63-70.
[3] BEARDON, A. F.: Iteration of Rational Functions. Springer-Verlag, New York, 1991.
[4] CARATHÉODORY, C.: Theory of Functions of a Complex Variable. Vol. 1., Vol. 2. Translated by F. Steinhardt. Chelsea Publishing Co., New York, N. Y. 1954.
[5] DOWNS, T. D.: Cauchy families of directional distributions closed under location and scale transformations, Open Stat. Prob. J. 1 (2009), no. 1, 76-92.
[6] DOWNS, T. D.-DOWNS, K. J.: Linear and directional domains with Cauchy probability distributions, Open Stat. Prob. J. 4 (2012), no. 1, 5-10.
[7] FIORENZA, A. -VINCENZI, G.: Limit of ratio of consecutive terms for general order$k$ linear homogeneous recurrences with constant coefficients, Chaos Solitons Fractals 44 (2011), no. 1-3, 145-152.
[8] GOLDSTERN, M.-TICHY, R. F.-TURNWALD, G.: Distribution of the ratios of the terms of a linear recurrence, Monatsh. Math. 107 (1989), no. 1, 35-55.
[9] GOLZY, M.-MARKATOU, M.: Poisson kernel-based clustering on the sphere: convergence properties, identifiability, and a method of sampling, J. Comput. Graph. Statist. 29 (2020), no. 4 758-770.
[10] JAMMALAMADAKA, S. R.-SENGUPTA, A.: Topics in Circular Statistics. Series on Multivariate Analysis, Vol 5. World Scientific Publishing Co., Inc., River Edge, NJ, 2001.
[11] KATO, S.-MCCULLAGH, P.: Some properties of a Cauchy family on the sphere derived from the Möbius transformations, Bernoulli 26 (2020), no. 4, 3224-3248.
[12] KATO, S.-SHIMIZU, K.-SHIEH, G.: A circular-circular regression model, Statist. Sinica 18 (2008), no. 2, 633-645.
[13] KATOK, A.-HASSELBLATT, B.: Introduction to the Modern Theory of Dynamical Systems. With a supplementary chapter by Katok and Leonardo Mendoza. Encyclopedia of Mathematics and its Applications, Vol. 54. Cambridge University Press, Cambridge, 1995.
[14] KELLEY, W. G.-PETERSON, A. C.: Difference Equations: An Introduction with Applications. Second edition. Harcourt/Academic Press, San Diego, California, 2001.
[15] KISS, P.: A Diophantine approximative property of the second-order linear recurrences, Period. Math. Hungar. 11 (1980), no. 4, 281-287.
[16] KISS, P.: Results on the ratios of the terms of second-order linear recurrences, Mathematica Slovaca 41 (1991), no. 3, 257-260.
[17] KISS, P.: An approximation problem concerning linear recurrences, In: Number Theory (Eger, 1996), De Gruyter, Berlin, 1998, pp. 289-293.
[18] KISS, P.-SINKA, Z.: On the ratios of the terms of second-order linear recurrences, Period. Math. Hungar. 23 (1991), no. 2, 139-143.

## DANIEL BEREND - RISHI KUMAR

[19] KISS, P.-TICHY, R. F.: Distribution of the ratios of the terms of a second-order linear recurrence, Nederl. Akad. Wetensch. Indag. Math. 48 (1986), no. 1, 79-86.
[20] KISS, P.-TICHY, R. F.: A discrepancy problem with applications to linear recurrences I, Proc. Japan Acad. Ser. A Math. Sci. 65 (1989), no. 5, 135-138.
[21] KISS, P.-TICHY, R. F.: A discrepancy problem with applications to linear recurrences II, Proc. Japan Acad. Ser. A Math. Sci., 65 (1989), no. 6, 191-194.
[22] KOSHY, T.: Fibonacci and Lucas Numbers with Applications. John Wiley \& Sons, New York, 2001.
[23] KUIPERS, L.-NIEDERREITER, H.: Uniform Distribution of Sequences. John Wiley, New York, 1974.
[24] LÉVY, P.: L’addition des variables aléatoires définies sur une circonférence, Bull. Soc. Math. France 67 (1939), 1-41.
[25] MARDIA, K. V.: Statistics of Directional Data, Academic Press, New York, 1972.
[26] MCCULLAGH, P.: Möbius transformation and Cauchy parameter estimation, Ann. Statist. 24 (1996), no. 2, 787-808.
[27] POINCARÉ, H.: Sur les équations linéaires aux différentielles ordinaires et aux différences finies, Amer. J. Math. 7 (1885), no. 3, 203-258.
[28] RAVINDRAN, P.-GHOSH, S.: Bayesian analysis of circular data using wrapped distributions, J. Stat. Theory Pract. 5 (2011), no. 4, 547-561.
[29] ROSS, S. M.: A First Course in Probability. Pearson Education Limited, Upper Saddle River, New Jersey, 2010.
[30] RUDIN, W.: Real and Complex Analysis. McGraw-HiII, New York, 1987.
[31] VIANA, W.-OLIVEIRA, K.: Foundations of Ergodic Theory. Cambridge University Press, Cambridge, 2016.
[32] WALCK, C.: Hand-book on Statistical Distributions for Experimentalists. Internal Report SUF-PFY/96-01 (last modification 10 Sept. 2007), Fysikum, University of Stockholm, Particle Physics Group, 2007.
[33] WALTERS, P.: An Introduction to Ergodic Theory. Springer-Verlag, Berlin, 1982.
[34] WINTNER, A.: On the shape of the angular case of Cauchy's distribution curves, Annals Math. Statist. 18 (1947), no. 4, 589-593.

Received February 22,2022
Accepted April 21,2022

## Daniel Berend

Departments of Mathematics and Computer Science
Ben-Gurion University of the Negev
Beer Sheva 84105
ISRAEL
E-mail: berend@math.bgu.ac.il

## Rishi Kumar

Department of Mathematics
Ben-Gurion University of the Negev
Beer Sheva 84105
ISRAEL
E-mail: kumarr@post.bgu.ac.il


[^0]:    Möbius transformation, Cauchy distribution, circular Cauchy distribution, unique ergodicity.
    (C) 2022 BOKU-University of Natural Resources and Life Sciences and Mathematical Institute, Slovak Academy of Sciences.
    2020 Mathematics Subject Classification: 11B37.
    The first author is supported in part by the Milken Families Foundation Chair in Mathematics. ©()®® Licensed under the Creative Commons BY-NC-ND 4.0 International Public License.

