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The primary function of THE FIBONACCI QUARTERLY is to serve as a focal point for widespread interest in the Fibonacci and related numbers, especially with respect to new results, research proposals, challenging problems, and innovative proofs of old ideas.

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# Gie Fibonacci Quarterly 

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## ON A NEW KIND OF NUMBERS

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## 1. Introduction

Recently, properties of the sequence $\left\{L_{2 n+1}\right\}$, where $L_{2 n+1}$ is a Lucas number of order $2 n+1$, were studied [1]. In the present paper, we introduce a new class of numbers defined by
(1.1) $f(n, k)=(-1)^{n-k}\binom{2 n+1}{n-k} L_{2 k+1}$,
where $n$ is any nonnegative integer and $0 \leq k \leq n$.
These numbers have the interesting property (see [1], (1.5)):
(1.2) $\sum_{k=0}^{n} f(n, k)=1$,
for every nonnegative integral value of $n$. Property (1.2) is very much analogous to the following property of Stirling numbers of the first kind (see [2], (6), p. 145):

$$
\begin{equation*}
\sum_{k=1}^{n} S_{n}^{k}=0 \tag{1.3}
\end{equation*}
$$

Also, these new numbers generalize the Catalan numbers in a nontrivial way. First, we recall that the Catalan numbers $c_{n}$ are defined by means of the generating relation ([5], p. 82)

$$
\begin{equation*}
\sum_{n=0}^{\infty} c_{n} t^{n}=(2 t)^{-1}(1-\sqrt{1-4 t}) \tag{1.4}
\end{equation*}
$$

or by the explicit formula ([5], p. 101)
(1.5) $\quad c_{n}=(n+1)^{-1}\binom{2 n}{n}$.

The following relationship is obvious:

$$
\text { (1.6) } \quad c_{n}=\frac{(-1)^{n} f(n, 0)}{(2 n+1)}
$$

Results obtained in this paper include a table, recurrence relations, generating functions, and summation formulas for these new numbers. In view of (1.6), many results reduce to their corresponding results for the Catalan numbers found in the literature. In our last section, we pose two significant open problems.

As usual $(\alpha)_{n}$ is Pochhammer's symbol and is defined by

$$
\text { (1.7) } \quad(\alpha)_{n}= \begin{cases}1 & \text { if } n=0 \\ \alpha(\alpha+1) \ldots(\alpha+n-1), & \text { for all } n \in\{1,2,3, \ldots\},\end{cases}
$$

${ }_{2} F_{1}$ will denote the hypergeometric function defined by
(1.8) $\quad{ }_{2} F_{1}\left[\begin{array}{c}a, b ; x \\ c ;\end{array}\right]=\sum_{n=0}^{\infty} \frac{(a)_{n}(b)_{n}}{(c)_{n}} \frac{x^{n}}{n!}, c \neq 0,-1,-2, \ldots$,
and the Jacobi polynomials are defined by

$$
\begin{align*}
& P_{n}^{(\alpha, \beta)}(x)=\frac{(1+\beta)_{n}}{n!}\left(\frac{x-1}{2}\right)^{n}{ }_{2} F_{1}\left[\begin{array}{l}
-n,-\alpha-n ; \frac{x+1}{x-1} \\
1+\beta ;
\end{array}\right] .  \tag{1.9}\\
& \text { 2. Table of } f(n, k)
\end{align*}
$$

In this section, we give a table of $f(n, k)$ produced by SCRATCHPAD-IBM's symbolic manipulation language.

TABLE OF $f(n, k)$

| $n^{k}$ | 0 | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 | 9 | 10 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1 |  |  |  |  |  |  |  |  |  |  |
| 1 | -3 | 4 |  |  |  |  |  |  |  |  |  |
| 2 | 10 | -20 | 11 |  |  |  |  |  |  |  |  |
| 3 | -35 | 84 | -77 | 29 |  |  |  |  |  |  |  |
| 4 | 126 | -336 | 396 | -261 | 76 |  |  |  |  |  |  |
| 5 | -462 | 1320 | -1815 | 1595 | -836 | 199 |  |  |  |  |  |
| 6 | 1716 | -5148 | 7865 | -8294 | 5928 | -2587 | 521 |  |  |  |  |
| 7 | -6435 | 20020 | -33033 | 39585 | -34580 | 20895 | -7815 | 1364 |  |  |  |
| 8 | 24310 | -77792 | 136136 | -179452 | 180880 | -135320 | 70856 | -23188 | 3571 |  |  |
| 9 | -92378 | 302328 | -554268 | 786828 | -883728 | 771324 | -504849 | 233244 | -67849 | 9349 |  |
| 10 | 352716 | -1175720 | 2238390 | -3372120 | 4124064 | -4049451 | 3118185 | -1814120 | 749910 | -196329 | 24476 |

3. Recurrence Relations

In equation (1.3) of [1], it is noted that

$$
\begin{equation*}
3 L_{2 n+1}-L_{2 n-1}=L_{2 n+3}, n \geq 2 . \tag{3.1}
\end{equation*}
$$

Using (3.1) with $n$ replaced by $n-1$ and (1.1), we see that

$$
\begin{align*}
(n & +k+1)(n+k) f(n, k)+3(n+k)(n-k+1) f(n, k-1)  \tag{3.2}\\
& +(n-k+1)(n-k+2) f(n, k-2)=0, k \geq 2 .
\end{align*}
$$

Furthermore, eliminating $L_{2 k+1}$ from

$$
f(n+1, k)=(-1)^{n+1-k}\binom{2 n+3}{n+1-k} L_{2 k+1}
$$

and

$$
f(n, k)=(-1)^{n-k}\binom{2 n+1}{n-k} L_{2 k+1}
$$

we obtain the formula
(3.3) $f(n+1, k)=\frac{-(2 n+3)(2 n+2)}{(n-k+1)(n+k+2)} f(n, k)$.

Following the method of proof of formula (3.3) and using (1.7), we can obtain its straightforward generalization in the form
(3.4) $f(n+m, k)=\frac{(-1)^{m}(2 n+2)_{2 m}}{(n-k+1)_{m}(n+k+2)_{m}} f(n, k)$,
where $m$ is a nonnegative integer.

## 4. Generating Relations

We first obtain generating functions for $f(n, k)$ with respect to $n$. That is, we prove the following theorem.

Theorem 4.1: Let $f(n, k)$ be defined by (1.1). Then

$$
\begin{equation*}
\sum_{n=0}^{\infty} \frac{f(n+k, k)}{2(k+n)+1} t^{n}=\frac{(1+v)^{2 k+1}}{2 k+1} L_{2 k+1} \tag{4.1}
\end{equation*}
$$

and

$$
\begin{equation*}
\sum_{n=0}^{\infty} f(n+k, k) t^{n}=\frac{(1+v)^{2 k+2}}{1-v} L_{2 k+1} \tag{4.2}
\end{equation*}
$$

where $v$ is the function of $t$ defined by
(4.3) $\quad v=-t(1+v)^{2}$.

Remark: In view of (1.6), (4.1) yields (1.4) while (4.2) yields the following generating relation for the Catalan numbers:

$$
\begin{equation*}
\sum_{n=0}^{\infty}(2 n+1) c_{n} t^{n}=\frac{(1+v)^{2}}{1-v}, v=t(1+v)^{2} \tag{4.4}
\end{equation*}
$$

Proof of (4.1) : First, multiply both sides of (1.1) by $(2 k+1) /[2(k+n)+1]$. Then sum over $n$ from 0 to $\infty$. Finally, appeal to the well-known identity ([4], p. 348, Prob. 212),

$$
\begin{equation*}
\sum_{n=0}^{\infty} \frac{\alpha}{\alpha+\beta n}\binom{\alpha+\beta n}{n} t^{n}=(1+u)^{\alpha}, u=t(1+u)^{\beta} \tag{4.5}
\end{equation*}
$$

to obtain (4.1).
Proof of (4.2) : Starting with the definition (1.1) of $f(n, k)$, we have

$$
\sum_{n=0}^{\infty} f(n+k, k) t^{n}=L_{2 k+1} \sum_{n=0}^{\infty}\binom{2 n+2 k+1}{n}(-t)^{n}=L_{2 k+1} \frac{(1+v)^{2 k+2}}{1-v}
$$

by virtue of the identity (see [4], p. 349, Prob. 216), which is

$$
\begin{align*}
\sum_{n=0}^{\infty}\binom{\alpha+(\beta+1) n}{n} t^{n} & =\frac{(1+u)^{\alpha+1}}{1-\beta u}  \tag{4.6}\\
u & =t(1+u)^{\beta+1}
\end{align*}
$$

Next, we prove the following theorem on generating functions involving double series:
Theorem 4.2: Let $f(n, k)$ be defined by (1.1). Then

$$
\begin{equation*}
\sum_{n=0}^{\infty} \sum_{k=0}^{\infty} f(n+k, k) t^{n+k}=(1-t)^{-1},|t|<1, \tag{4.7}
\end{equation*}
$$

and
(4.8) $\sum_{n=0}^{\infty} \sum_{k=0}^{\infty} \frac{f(n+k, k)}{2(k+n)+1} t^{n+k}=(1+v) F_{4}\left[1, \frac{1}{2} ; \frac{3}{2}, \frac{1}{2} ; \frac{-v}{4}, \frac{-5 v}{4}\right]$,
where $F_{4}$ is Appell's double hypergeometric function of the fourth kind defined by ([6], p. 14), that is,

$$
\text { (4.9) } F_{4}\left[a, b ; c, c^{\prime} ; x, y\right]=\sum_{m=0}^{\infty} \sum_{n=0}^{\infty} \frac{(a)_{m+n}(b)_{m+n}}{(c)_{m}\left(c^{\prime}\right)_{m}} \frac{x^{m}}{m!} \frac{y^{n}}{n!}, \sqrt{|x|}+\sqrt{|y|}<1 .
$$

Proof of (4.7): By making use of (1.1), we observe that

$$
\sum_{n=0}^{\infty} \sum_{k=0}^{\infty} f(n+k, k) t^{n+k}=\sum_{k=0}^{\infty} L_{2 k+1} t^{k} \sum_{n=0}^{\infty}\binom{2 n+2 k+1}{n}(-t)^{n} .
$$

Summing the inner series with the help of (4.6) and then interpreting the resulting expression by means of the generating relation (1.4) of [1], which is,
(4.10) $\sum_{n=0}^{\infty} L_{2 n+1} t^{n}=(1+t)\left(1-3 t+t^{2}\right)^{-1},|t|<1$,
we are led to (4.7). Alternatively, (4.7) can be obtained by using (1.2).
Proof of (4.8) : Comparing (4.10) with the following generating function for Jacobi polynomials (see [3], Eq. 10, p. 256),
(4.11) $\sum_{n=0}^{\infty} \frac{(\alpha+\beta+1)_{n} P_{n}^{(\alpha, \beta)}(x)}{(1+\alpha)_{n}} t^{n}$

$$
=(1-t)^{-1-\alpha-\beta}{ }_{2} F_{I}\left[\begin{array}{l}
\frac{1}{2}(1+\alpha+\beta), \frac{1}{2}(2+\alpha+\beta) ; \frac{2 t(x-1)}{(1-t)^{2}} \\
1+\alpha ;
\end{array}\right.
$$

we obtain the formula
(4.12) $L_{2 n+1}=\frac{n!}{(1 / 2)_{n}} P_{n}^{(1 / 2,-1 / 2)}(3 / 2)$.

If we use (4.12) in the following generating function (see [3], p. 271),
(4.13) $\sum_{n=0}^{\infty} \frac{(\gamma)_{n}(\delta)_{n} P_{n}^{(\alpha, \beta)}(x)}{(1+\alpha)_{n}(1+\beta)_{n}} t^{n}=F_{4}\left[\gamma, \delta ; 1+\alpha, 1+\beta ; \frac{1}{2} t(x-1), \frac{1}{2} t(x+1)\right]$,
remembering that
(4.14) $(3 / 2)_{n}=(1 / 2)_{n}(2 n+1)$,
we find that $L_{2 n+1}$ satisfies the generating relation
(4.15) $\sum_{n=0}^{\infty} \frac{L_{2 n+1}}{2 n+1} t^{n}=F_{4}\left(1, \frac{1}{2} ; \frac{3}{2}, \frac{1}{2} ; \frac{t}{4}, \frac{5 t}{4}\right)$.

If we now start with the left-hand side of (4.8), we have

$$
\sum_{n, k=0}^{\infty} \frac{f(n+k, k)}{2(n+k)+1} t^{n+k}=\sum_{k=0}^{\infty} \frac{L_{2 k+1}}{2 k+1} t^{k} \sum_{n=0}^{\infty} \frac{2 k+1}{2(k+n)+1}\binom{2 n+2 k+1}{n}(-t)^{n} .
$$

Summing the inner series by using (4.5), we get

$$
\sum_{n, k=0}^{\infty} \frac{f(n+k, k) t^{n+k}}{2(n+k)+1}=(1+v) \sum_{k=0}^{\infty} \frac{L_{2 k+1}}{2 k+1}(-v)^{k} .
$$

Interpreting the last infinite series by means of (4.15) along with the second member of the generating function, (4.8) follows at once. This concludes the proof of Theorem 4.2.

## 5. Summation Formulas

In this section we propose to prove the following summation formulas:

$$
\begin{equation*}
\sum_{m=0}^{n-1}\left\{f(n+k-m-1, k)+\frac{L_{2 k+1}}{L_{2 k-1}} f(n+k-m-1, k-1)\right\} f(m, 0) \tag{5.1}
\end{equation*}
$$

$$
=f(n+k, k)-f(n, 0) L_{2 n+1}, k \geq 1
$$

$$
\begin{equation*}
\sum_{m=0}^{n-1} \frac{f(m, 0) f(n-m-1,0)}{(2 m+1)\{2(n-m)-1\}}=-\frac{f(n, 0)}{2 n+1} . \tag{5.2}
\end{equation*}
$$

(5.3) $\sum_{m=0}^{\infty} \frac{\{n-2 m(1+k)\}}{(2 m+1)\{2(k+n-m)+1\}} f(n+k-m, k) f(m, 0)=0$.

$$
\begin{align*}
& \sum_{m=0}^{n-1} f(m, 0)\left[\frac{f(n-m+k-1, k)}{2(k+n-m)-1}+\frac{L_{2 k+1}}{L_{2 k-1}} \frac{f(n+k-m-1, k-1)}{(2 m+1)(2 k+1)}\right]  \tag{5.4}\\
& =\frac{f(n+k, k)}{2(k+n)+1}-\frac{L_{2 k+1}}{(2 k+1)} \frac{f(n, 0)}{(2 n+1)}, k \geq 1 .
\end{align*}
$$

$$
\begin{align*}
& \text { (5.5) } \sum_{k=0}^{n}(-1)^{k}\left[\frac{f(n, n-k)}{L_{2(n-k)+1}}+\frac{f(n, n-k+1)}{L_{2(n-k)}+3}\right]=(2 n+1) c_{n}  \tag{5.5}\\
& \text { (5.6) } \quad \sum_{k=0}^{n} \frac{f(n, k) f(n, n-k)}{L_{2 k+1} L_{2(n-k)+1}}=\frac{f(2 n, n)}{L_{2 n+1}}-\frac{f(2 n, n+1)}{L_{2 n+3}}
\end{align*}
$$

Remark: In view of relationship (1.6), (5.2) and (5.3) yield the following formulas for the Catalan numbers:
(5.7) $\sum_{m=0}^{n} c_{m} c_{n-m}=c_{n+1}$,
and

$$
\begin{equation*}
n \sum_{m=0}^{\infty} c_{m} c_{n-m}=2 \sum_{m=0}^{\infty} m c_{m} c_{n-m}, \tag{5.8}
\end{equation*}
$$

respectively.
Proof of (5.1): Changing the dummy index $k$ to $k-1$ in (4.2), we get
(5.9) $\sum_{n=0}^{\infty} f(n+k-1, k-1) t^{n}=L_{2 k-1} \frac{(1+v)^{2 k}}{1-v}$.

On the other hand, for $k=0$, (4.2) reduces to
(5.10) $\sum_{n=0}^{\infty} f(n, 0) t^{n}=\frac{(1+v)^{2}}{1-v}$.

In view of (5.9), (4.2) can be written as

$$
\sum_{n=0}^{\infty} f(n+k, k) t^{n}=\frac{L_{2 k+1}}{L_{2 k-1}}(1+v)^{2} \sum_{n=0}^{\infty} f(n+k-1, k-1)
$$

which can be rewritten in the form

$$
\begin{aligned}
& {\left[1-t \sum_{m=0}^{\infty} f(m, 0) t^{m}\right] \sum_{n=0}^{\infty} f(n+k, k) t^{n}} \\
& =\frac{L_{2 k+1}}{L_{2 k-1}} \sum_{n=0}^{\infty} f(m, 0) f(n+k-1, k-1) t^{n+m}
\end{aligned}
$$

by virtue of (5.10).
Equating the coefficients of $t^{n}$ and using (2.4), we arrive at (5.1).
Proof of (5.2): Setting $k=0$ in (4.1), we get
(5.11) $\sum_{n=0}^{\infty} \frac{f(n, 0)}{2 n+1} t^{n}=1+v$.

In view of the definition of $v$ in (4.3), (5.11) can be written as

$$
\sum_{n=0}^{\infty} \frac{f(n, 0)}{2 n+1} t^{n}=1-t(1+v)^{2}=1-t \sum_{n, m=0}^{\infty} \frac{f(n, 0) f(m, 0)}{(2 n+1)(2 m+1)} t^{n+m}
$$

Comparing coefficients of $t^{n}$, we get (5.2).
Proof of (5.3) : Combining (4.1) and (4.2), we have

$$
\sum_{n=0}^{\infty} f(n+k, k) t^{n}=\frac{(2 k+1)(1+v)}{1-v} \sum_{n=0}^{\infty} \frac{f(n+k, k)}{2(k+n)+1} t^{n}
$$

Multiplying both sides by $(1+v)$ and then using (5.10) and (5.11), we obtain

$$
\sum_{n, m=0}^{\infty} \frac{f(m, 0) f(n+k, k)}{2 m+1} t^{n+m}=(2 k+1) \sum_{n, m=0}^{\infty} \frac{f(m, 0) f(n+k, k)}{2(k+n)+1} t^{n+m} .
$$

By equating the coefficients of $t^{n}$, we get (5.3).

Proof of (5.4): This proof is similar to that of (5.1) and, hence, will be omitted.
Proof of (5.5) ; Equation (5.5) is an immediate consequence of (1.1), the following identity (see [5], p. 65),
(5.12) $\binom{2 n+1}{k}=\sum_{k=0}^{n}\left[\binom{2 n+1}{k}-\binom{2 n+1}{k-1}\right]$,
and (1.5).
Proof of (5.6): Using (1.1), we have

$$
\sum_{k=0}^{n} \frac{f(n, k) f(n, n-k)}{L_{2 k+1} L_{2(n-k)}+1}=(-1)^{k} \sum_{k=0}^{n}\binom{2 n+1}{k}\binom{2 n+1}{n-k}=(-1)^{k}\binom{4 n+2}{n}
$$

where we obtained the last equation by using the Vandermonde addition formula, which is
(5.13) $\sum_{k=0}^{n}\binom{x}{k}\binom{y}{n-k}=\binom{x+y}{n}$.

Appealing to the binomial identity

$$
\begin{equation*}
\binom{x+1}{n}=\binom{x}{n}+\binom{x}{n-1} \tag{5.14}
\end{equation*}
$$

we have

$$
\sum_{k=0}^{n} \frac{f(n, k) f(n, n-k)}{L_{2 k+1} L_{2(n-k)}+1}=(-1)^{n}\left\{\binom{4 n+1}{n}+\binom{4 n+1}{n-1}\right\}
$$

Using (1.1) to interpret the right-hand side, we arrive at (5.6).

## 6. Questions

The most obvious questions arising from this work are:
(i) Do the numbers $f(n, k)$ have a nice combinatorial meaning?
(ii) We have seen that $f(n, k)$ has a property analogous to the Stirling numbers of the first kind and that they also generalize the Catalan numbers. Is it possible to associate $f(n, k)$ with some other known mathematical objects?

## Acknowledgment

I wish, to thank the referee for his many helpful suggestions.

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# SOME SEQUENCES OF LARGE INTEGERS 

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## 1. Introduction

One of the many interesting problems posed in the book Unsolved Problems in Number Theory [1] concerns the sequence

$$
n x_{n}=x_{n-1}^{m}\left(x_{n-1}+n-1\right), x_{1} \in N .
$$

It was introduced by Fritz Göbel and has been studied by Lenstra [1] for $m=1$ and $x_{1}=2$. Lenstra states that $x_{n}$ is an integer for all $n \leq 42$, but $x_{43}$ is not. For $m=2$ and $x_{1}=2$, David Boyd and Alf van der Poorten state that for $n \leq 88$ the only possible denominators in $x_{n}$ are products of powers of $2,3,5$, and 7. Why do these denominators cause a problem? Is it possible to find even longer sequences of integers by choosing different values of $x_{1}$ and $m$ ? These questions were posed by M. Mudge [2].

The terms in these sequences grow fast. For $m=1, x_{1}=2$, the first ten terms are:

3, 5, 10, 28, 154, 3520, 15518880, 267593772160, 160642690122633501504.
If the number of digits in $x_{n}$ is denoted $N(n)$, then $N(11)=43, N(12)=85$, $N(13)=168, N(14)=334, N(15)=667, N(16)=1332$, and $N(17)=2661$. The last integer in this sequence, $x_{42}$, has approximately 89288343500 digits.

The purpose of this study is to find a method of determining the number of integers in the sequence and apply the method for the parameters $1 \leq m \leq 10$ and $2 \leq x_{1} \leq 11$. In particular, the problem of Boyd and van der Poorten will be solved. Some explanations will be given to why some of these sequences are so long. It will be observed and explained why the integer sequences are in general longer for even than for odd values of $m$.

## 2. Method

For given values of $x_{1}$ and $m$ consider the equation

$$
\begin{equation*}
k x_{k}=x_{k-1}\left(x_{k-1}^{m}+k-1\right) \tag{1}
\end{equation*}
$$

where the prime factorization of $k$ is given by

$$
\begin{equation*}
k=\prod_{i=1}^{l} p_{i}^{n_{i}} . \tag{2}
\end{equation*}
$$

Let us assume that $x_{k-1}$ is an integer and expand $x_{k-1}$ and $x_{k-1}^{m}+k-1$ in a number system with $G_{i}=p_{i}^{t_{i}},\left(t_{i}>n_{i}\right)$ as base.

$$
\begin{equation*}
x_{k-1}=\sum_{j} \alpha_{j} G_{i}^{j} \quad\left(0 \leq \alpha_{i}<G_{i}\right) \tag{3a}
\end{equation*}
$$

and

$$
\begin{equation*}
x_{k-1}^{m}+k-1=\sum_{j} b_{j} G_{i}^{j} \quad\left(0 \leq b_{j}<G_{i}\right) . \tag{3b}
\end{equation*}
$$

Since $x_{k-1} \neq 0$, it is always possible to choose $t_{i}$ so that $\alpha_{0} \neq 0$ and $b_{0} \neq 0$. With this $t_{i}$ we have

$$
\begin{equation*}
x_{k-1}\left(x_{k-1}^{m}+k-1\right)=\sum_{j, \ell} \alpha_{j} b_{\ell} G_{i}^{j+\ell} \equiv a_{0} b_{0}\left(\bmod G_{i}\right) . \tag{4}
\end{equation*}
$$

The congruence

$$
\begin{equation*}
k x_{k} \equiv \alpha_{0} b_{0}\left(\bmod G_{i}\right) \tag{5}
\end{equation*}
$$

is soluble iff $\left(k, G_{i}\right) \mid a_{0} b_{0}$, or, in this case, iff $p_{i}^{n_{i}} \mid \alpha_{0} b_{0}$. But, if $p_{i}^{n_{i}} \mid \alpha_{0} b_{0}$, then by (4) we also have

$$
p_{i}^{n_{i}} \mid x_{k-1}\left(x_{k-1}^{m}+k-1\right)
$$

Furthermore, if (5) is soluble for all expansions originating from (2), then it follows that

$$
k \mid x_{k-1}\left(x_{k-1}^{m}+k-1\right)
$$

and, consequently, that $x_{k}$ is an integer. The solution $x_{k}\left(\bmod G_{i}\right)$ to $k x_{k} \equiv$ $a_{0} b_{0}\left(\bmod G_{i}\right)$ is equal to the first term in the expansion of $x_{k}$ using the equivalent of (3a). The previous procedure is repeated using (3b), (4), and (5) to examine if $x_{k+1}\left(\bmod G_{i}\right)$ is an integer.

From the computational point of view, the testing is done up to a certain pre-set limit $k=k_{\max }$ for consecutive primes $p=2,3,5,7, \ldots$ to $p \leq k_{\max }$. One of three things will happen:

1. All congruences are soluble modulus $G_{i}$ for $k \leq k_{\max }$ for all $p_{i} \leq k_{\max }$.
2. $a_{0} b_{0}=0$ for a certain set of values $k \leq k_{\max }, p_{i} \leq k_{\max }$.
3. The congruence $k x_{k} \equiv \alpha_{0} b_{0}\left(\bmod G_{i}\right)$ is soluble for all $k<n \leq k_{\max }$, but not soluble for $k=n$ and $p=p_{i}$.
In cases 1 and 2 increase $k_{\max }$, respectively, $t_{i}$ in $G_{i}=p_{i}^{t_{i}}$ (if computer facilities permit) and recalculate. In case $3, x_{n}$ is not an integer, viz. $n$ has been found so that $x_{k}$ is an integer for $k<n$ but not for $k=n$.

## 3. Results

The results from using this method in the 100 cases $1 \leq m \leq 10,2 \leq x_{1} \leq 11$ are shown in Table 1. In particular, it shows that the integer sequence holds up to $n=88$ for $m=2, x_{1}=2$ which corresponds to the problem of Boyd and van der Poorten. The longest sequence of integers was found for $x_{1}=11, m=2$. For these parameters, the 600 first terms are integers, but $x_{601}$ is not. In the 100 cases studied, only 32 different primes occur in the terminating values $n$. In 7 cases, the integer sequences are broken by values of $n$ which are not primes. In 6 of these, the value of $n$ is 2 times a prime which had terminated other sequences. For $x_{1}=3, m=10$, the sequence is terminated by $n=2 * 13^{2}$. The prime 239 is involved in terminating 10 of the 100 sequences studied. It occurs 3 times for $m=6$ and 7 times for $m=10$. It is seen from the table that integer sequences are in general longer for even than for odd values of $m$.

TABLE 1. $x_{n}$ is the first noninteger term in the sequence defined by $n x_{n}=x_{n-1}\left(x_{n-1}^{m}+n-1\right)$. The table gives $n$ for parameters $x_{1}$ and $m$.

| $m$ | $x_{1}=2$ | $x_{1}=3$ | $x_{1}=4$ | $x_{1}=5$ | $x_{1}=6$ | $x_{1}=7$ | $x_{1}=8$ | $x_{1}=9$ | $x_{1}=10$ | $x_{1}=11$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 1 | 43 | 7 | 17 | 34 | 17 | 17 | 51 | 17 | 7 | 34 |
| 2 | 89 | 89 | 89 | 89 | 31 | 151 | 79 | 89 | 79 | 601 |
| 3 | 97 | 17 | 23 | 97 | 149 | 13 | 13 | 83 | 23 | 13 |
| 4 | 214 | 43 | 139 | 107 | 269 | 107 | 214 | 139 | 251 | 107 |
| 5 | 19 | 83 | 13 | 19 | 13 | 37 | 13 | 37 | 347 | 19 |
| 6 | 239 | 191 | 359 | 419 | 127 | 127 | 239 | 191 | 239 | 461 |
| 7 | 37 | 7 | 23 | 37 | 23 | 37 | 17 | 23 | 7 | 37 |
| 8 | 79 | 127 | 158 | 79 | 103 | 103 | 163 | 103 | 163 | 79 |
| 9 | 83 | 31 | 41 | 83 | 71 | 83 | 71 | 23 | 41 | 31 |
| 10 | 239 | 338 | 139 | 137 | 239 | 239 | 239 | 239 | 239 | 389 |

## 4. A Model To Explain Some Features of the Sequence

The congruence

$$
x(k) \equiv \alpha(k)(\bmod p), \alpha(k) \in\{-1,0,1, \ldots, p-2\}
$$

studied in a number system with a sufficiently large base $p^{t}$, is of particular interest when looking at the integer properties of the sequence. Five cases will be studied. These are:

1. $\alpha(k)$ does not belong to cases $2,3,4$, or 5 below
2. $\alpha(k)=-1, p \neq 2$
3. $\alpha(k)=0$
4. $\alpha(k)=1$
5. $\alpha(k)=\alpha(k+1)$ and/or $\alpha(k)=\alpha(k-1), \alpha(k) \neq-1,0,1$

These cases are mutually exclusive; however, in case 5 there may be more than one sequence of the described type for a given $p$, for example, for $m=10, x_{1}=$ 7 , and $p=11$, we have $\alpha(k)=7$ for $k=1,2, \ldots, 10$ and $\alpha(k)=4$ for $k=11$, $12, \ldots, 15$. Therefore, when running through the values of $k$ for a given $p$, it is possible to classify $\alpha(k)$ into states corresponding to cases $1,2,3,4$ or into one of several possible states corresponding to case 5 . In this model, $\alpha(1)$ appears as a result of creation rather than transition from one state to another but, formally, it will be considered as resulting from transition from a state $0(k=0)$ to the state corresponding to $\alpha(1)$.

The study of transitions from one state to another in the above model is useful in explaining why there are such long sequences of integers and why they are in general longer for even than for odd $m$. Table 2 shows the number of transitions of each kind in the 100 cases studied. Let $\alpha_{r}$ be the number of transitions from state $r$ to state $s$ :

$$
A_{r}=\sum_{r} a_{r s}, B_{s}=\sum_{s} a_{r s}, Q_{s}=100 A_{s} / B_{s} .
$$

(Note that $r$ and $s$ refer to states not rows and columns in Table 2.) The transitions for odd and even values of $m$ are treated separately. It is seen that transitions from states 4, 5, and 2 (for even $m$ ) are rare. Only between $5 \%$ and $14 \%$ of all such states "created" are "destroyed," while the corresponding percentage for other transitions range between $85 \%$ and $99 \%$. It is the fact that transitions from certain states are rare, which makes some of these integer sequences so long. That transitions from state 2 are rare for even $m$ (11\%) and frequent for odd $m$ (99\%) make the integer sequences in general longer for even than for odd $m$. In all the many transitions observed, it was noted that certain types (underscored in Table 2) only occurred for values of $k$ divisible by $p$, while other types never occurred for $k$ divisible by $p$. Transitions from state 3 all occur for $k$ divisible by $p$ but, unlike the other transitions which occur for $k$ divisible by $p$, they have a high frequency. Some of the observations made on the model are explained in the remainder of this paper.

TABLE 2. The number of transitions of each type for odd and even $m$

| From state | To state 1 |  | To state 2 |  | To state 3 |  | To state 4 |  | To state 5 |  | $A_{r}$ |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m | 21 m |
| 0 | 467 | 1847 | 38 | 40 | 60 | 60 | 55 | 55 | 32 | 69 | 652 | 2071 |
| 1 |  |  | 220 | 701 | 252 | 791 | 247 | 642 | 75 | 307 | 794 | 2241 |
| 2 | 181 | 55 |  |  | 71 | 21 | 39 | 7 | 2 | 0 | 293 | 83 |
| 3 | $\underline{202}$ | $6 \overline{34}$ | 36 | 30 |  |  | 111 | 80 | $\underline{9}$ | 16 | 358 | 760 |
| 4 | 20 | 35 | $\frac{2}{2}$ | 6 | 39 | 12 |  |  |  | $\underline{3}$ | 61 | 56 |
| 5 | $\underline{2}$ | $\underline{2}$ | $\underline{1}$ | $\underline{2}$ | 0 | $\underline{3}$ | 0 | $\underline{2}$ | $\underline{2}$ | 11 | 5 | 20 |
| $B_{s}$ | 872 | 2573 | 297 | 779 | 422 | 887 | 452 | 786 | 120 | 406 | 2163 | 5431 |
| $Q_{s} \%$ | 92 | 95 | 99 | 11 | 85 | 86 | 14 | 8 | 5 | 5 |  |  |

Transitions from state 4 and, for even $m$ only, from state 2
It is evident from $k x_{k}=x_{k-1}\left(x_{k-1}^{m}+k-1\right)$ that, if $x_{k-1} \equiv \pm 1(\bmod p)$ and $(k, p)=1$, then $x_{k}= \pm 1(\bmod p)$. Assume that we arrive at $x_{k-1} \equiv \pm 1(\bmod p)$ for $k<p-m$ and $m<p$. We can then write

$$
\begin{equation*}
x_{p-m-1} \equiv \pm 1+\alpha p\left(\bmod p^{2}\right), 0 \leq \alpha<p \tag{6}
\end{equation*}
$$

and
(7)

$$
x_{p-m-1}^{m} \equiv( \pm 1+\alpha p)^{m} \equiv 1 \pm m \alpha p\left(\bmod p^{2}\right) \quad(m \text { even }) .
$$

Equations (6) and (7) give

$$
(p-m) x_{p-m} \equiv \pm(p-m)\left(\bmod p^{2}\right)
$$

or, since $(p-m, p)=1$,

$$
x_{p-m} \equiv \pm 1\left(\bmod p^{2}\right) \text { or } x_{k} \equiv \pm 1\left(\bmod p^{2}\right) \text { for } p-m \leq k \leq p-1
$$

For $k=p$, we have

$$
p x_{p} \equiv \pm 1(1+p-1)\left(\bmod p^{2}\right)
$$

or, after division by $p$ throughout

$$
x_{p} \equiv \pm 1(\bmod p)
$$

It is now easy to see that $x_{k} \equiv 1(\bmod p)$ continues to hold also for $k>p$. The integer sequence may, however, be broken for $k=p^{2}$.
Transitions from state 3
Let us assume that $x_{j} \equiv 0(\bmod p)$ for some $j<p$. If $(j+1, p)=1$, it follows that $x_{j+1} \equiv 0(\bmod p)$ or, generally, $x_{k} \equiv 0(\bmod p)$ for $j \leq k \leq p-1$. For $k=p-1$, we can write $x_{p-1} \equiv p a\left(\bmod p^{2}\right), 0 \leq \alpha<p-1$. We then have
$p x_{p} \equiv p \alpha\left(p^{m} \alpha^{m}+p-1\right)\left(\bmod p^{2}\right)$,
from which follows $x_{p} \equiv-\alpha(\bmod p)$, viz. $x_{p}$ is an integer; however, if $a \nmid 0$, the state is changed.
Transitions from states of type 5
When, for some $j<p-1$, it happens that $x_{j}^{m} \equiv 1(\bmod p)$, it is easily seen that $x_{k} \equiv x_{j}(\bmod p)$ for $j \leq k<p$. This implies

$$
p x_{p} \equiv x_{j}(1+p-1) \quad(\bmod p)
$$

from which it is seen that $x_{p}$ may not be congruent to $x_{j}(\bmod p)$ but also that $x_{p}$ is an integer.

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## $* * * * *$

# CHARACTERISTICS AND THE THREE GAP THEOREM 

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## 1. Introduction

In order to determine the sequence $v=([m \alpha], m=1,2,3, \ldots)$, for irrational $\alpha$ (where $[x]$ denotes the largest integer not exceeding $x$ ), Bernoulli [1] considered the sequence of differences $d_{1}, d_{2}, d_{3}, \ldots$, where

$$
\begin{equation*}
d_{m}=[(m+1) \alpha]-[m \alpha], m=1,2,3, \ldots \tag{1}
\end{equation*}
$$

Clearly then,

$$
[m \alpha]=\sum_{i=1}^{m-1} d_{i}+[\alpha], m=3,4,5, \ldots .
$$

Thus, knowing the first two terms of $v$, one can then determine the entire sequence from (1). For example, with $\alpha=\sqrt{2}$, we have the following.

| $m$ | $d_{m}$ | $[m \alpha]$ |
| :---: | :---: | :---: |
| 1 | 1 | 1 |
| 2 | 2 | 2 |
| 3 | 1 | 4 |
| 4 | 2 | 5 |
| 5 | 1 | 7 |
| 6 | 1 | 8 |
| 7 | 2 | 9 |
| 8 | 1 | 11 |
| 9 | 2 | 12 |
| 10 | 1 | 14 |

It may be shown that $d_{m}$ may only equal $[\alpha]$ or $[\alpha]+1$ (that is, 0 or 1 when $0<\alpha<1$ ). If we replace $[\alpha]$ by $s$ (small) and $[\alpha]+1$ by 2 (large), then we obtain a string of such characters. This we will refer to as the characteristic of $\alpha$. For example, the characteristic of $\alpha=\sqrt{2}$ is sslslsslsl... .

String operations may be used to generate the characteristic from its first few terms, by utilizing the continued fraction expansion of $\alpha$. Bernoulli was the first to guess the rules which were the basis of these string operations. These were reformulated in a more attractive form by Christoffel [2]. However, it had to wait until Markoff [9] before the first proofs were offered. In Section 4 we show how the characteristic is generated.

In this paper we demonstrate a rather intriguing connection between the characteristic of $\alpha$ and the sequence of arcs or gaps formed by the partition of the circle by the successive placement of points by the angle $\alpha$ revolutions. The connection is not immediately obvious and does not hold for all values of $\alpha$. We use results from the Three Gap Theorem, a result first conjectured by Steinhaus (see $[6,10,11,13-15,18,19]$ ) which states that $N$ points placed on the circle as above partition it into gaps of either three or two different lengths.

Consider such a circle when $N$ is equal to the denominator of a convergent [see (2)] to $\alpha$. Only in this case is the circle partitioned into gaps of exactly two different lengths. We can label these gaps as large or small, assigning $Z$ or $s$ where appropriate and thus we have a string of gap types,
ordered clockwise about the circle, with the first element describing the gap adjacent to the origin.

We show that when $N$ is the denominator of a total convergent [see (3)] this string, after a trivial permutation, forms the first few terms of the characteristic, but only for special values of $\alpha$ (for example, those numbers with identical terms in their continued fraction expansion). One such value is the golden number, $\alpha=\tau=(\sqrt{5}-1) / 2$. The golden number's characteristic has interesting properties (see [16]) and we give it a special name-the Golden Sequence.

In order to state Christoffel's rule for generating the characteristic, we introduce in Section 2 some aspects from the theory of continued fractions. The Three Gap Theorem is later described in more detail in Section 3 before we prove our main result (in Section 4.2).

## 2. Continued Fractions

$$
\begin{aligned}
& \text { Write } t_{0}=\alpha \text { and express (for } n=0,1,2, \ldots \text { ), } \\
& \qquad \begin{aligned}
\alpha_{n} & =\left[t_{n}\right] \\
t_{n+1} & =\frac{1}{\left\{t_{n}\right\}},
\end{aligned}
\end{aligned}
$$

where $\{x\}=x-[x]$ is the fractional part of $x$. Thus, we can generate the simple continued fraction expansion of $\alpha$, namely,

$$
\begin{aligned}
\alpha & =a_{0}+\frac{1}{a_{1}+\frac{1}{a_{2}+\frac{1}{a_{3}+\cdots}}} \\
& =\left\{a_{0} ; a_{1}, a_{2}, a_{3}, \ldots\right\}
\end{aligned}
$$

The partial convergents to $\alpha$ are defined as

$$
\begin{equation*}
\frac{p_{n, i}}{q_{n, i}}=\left\{a_{0} ; a_{1}, a_{2}, \ldots, a_{n-1}, i\right\}, \quad i=1,2, \ldots, a_{n}-1 \tag{2}
\end{equation*}
$$

while

$$
\begin{equation*}
\frac{p_{n, a_{n}}}{q_{n, a_{n}}}=\frac{p_{n}}{q_{n}}=\left\{a_{0} ; a_{1}, a_{2}, \ldots, a_{n-1}, a_{n}\right\} \tag{3}
\end{equation*}
$$

defines the total convergents.
For example, the continued fraction of $\tau$ is given by

$$
\tau=\{0 ; 1+\tau\}=\{0 ; 1,1+\tau\}=\{0 ; 1,1,1, \ldots\}
$$

A11 convergents to $\tau$ are total convergents and

$$
p_{n}=q_{n-1}=F_{n}=F_{n-1}+F_{n-2}, n \geq 1, F_{-1}=1, F_{0}=0
$$

We quote some results from the theory of continued fractions (see Khintchine [7]);

$$
\begin{align*}
& \frac{p_{n, i}}{q_{n, i}}=\frac{p_{n-2}+i p_{n-1}}{q_{n-2}+i q_{n-1}}, p_{-2}=q_{-1}=0, q_{-2}=p_{-1}=1,  \tag{4}\\
& p_{n-1} q_{n, i}-q_{n-1} p_{n, i}=(-1)^{n} \\
& q_{n} \alpha-p_{n}=\frac{(-1)^{n}}{t_{n+1} p_{n}+p_{n+1}}, \\
& q_{n, i}\left\|q_{n-1} \alpha\right\|+q_{n-1}\left\|q_{n, i}\right\|=1 \\
& p_{n, i}\left\|q_{n-1} \alpha\right\|+p_{n-1}\left\|q_{n, i} \alpha\right\|=\alpha
\end{align*}
$$

$$
\begin{align*}
& \min _{0<q<q_{n+1}}\|q \alpha\|=\left\|q_{n} \alpha\right\|,  \tag{9}\\
& \left\|q_{n} \alpha\right\|= \begin{cases}1-\left\{q_{n} \alpha\right\}, & n \text { odd } \\
\left\{q_{n} \alpha\right\}, & n \text { even }\end{cases} \tag{10}
\end{align*}
$$

where $\|q \alpha\|=|q \alpha-p|, p=[q \alpha+1 / 2]$. That is, $\|q \alpha\|$ is equal to the absolute difference between $q \alpha$ and its nearest integer. Note that $p_{n}=\left[q_{n} \alpha+1 / 2\right]$.

Also, if $\alpha=\{0 ; \alpha, \alpha, \ldots\}=\left(\sqrt{\alpha^{2}+4}-\alpha\right) / 2$, then

$$
\begin{equation*}
p_{n}=\frac{(1 / \alpha)^{n}-(-\alpha)^{n}}{\alpha+1 / \alpha}=q_{n-1} . \tag{11}
\end{equation*}
$$

## 3. The Three Gap Theorem

The reader is referred to van Ravenstein [18] for an account of the Three Gap Theorem as well as the proofs of many of the results used in this section. Alternatively, the reader may see van Ravenstein [19] where the theorem is also discussed with special reference to the golden number.

### 3.1 Order of Points

Consider $N$ points placed in succession on a circle at an angle of $\alpha$. We are interested in determining the order of the points as they appear in clockwise order on the circle. This is equivalent to ordering (\{n $\}=n \alpha \bmod 1$, $n=0,1,2, \ldots, N-1)$ into an ascending sequence. ( $y \bmod x=y-x[y / x]=$ $x\{y / x\}$.) Let $\left(\left\{u_{j} \alpha\right\}\right), j=1,2, \ldots, N$ be that ordered sequence. That is,

$$
\left\{u_{1}, u_{2}, \ldots, u_{N}\right\}=\{0,1, \ldots, N-1\},
$$

where $\left\{u_{j} \alpha\right\}<\left\{u_{j+1} \alpha\right\}$. It is shown in Slater [11] and Sós [14] (or see [18], Th. 2.2) that the elements $u_{j}$ are obtained by the following relation,

$$
u_{j+1}-u_{j}= \begin{cases}u_{2}, & 0 \leq u_{j}<N-u_{2}  \tag{12}\\ u_{2}-u_{N}, & N-u_{2} \leq u_{j}<u_{N} \\ -u_{N}, & u_{N} \leq u_{j}<N\end{cases}
$$

for $j=1,2, \ldots, N, u_{1}=u_{N+1}=0$. Points $u_{j}$ and $u_{j+1}$ delimit the $j$ th gap, which is of length $\left\{\left(u_{j+1}-u_{j}\right) \alpha\right\}$.

Here, we will only be concerned with the case where the circle is partitioned into gaps of just two different lengths. This occurs when $N=u_{2}+u_{N}$ or, equivalently, when $N$ is the denominator of a convergent to $\alpha$.

It may be shown (from [18], Lemma 2.1) that, for $N=u_{2}+u_{N}=q_{n, i}(i=1$, $2, \ldots, a_{n}, n \geq 2$ ),

$$
\begin{equation*}
u_{j}=\left((-1)^{n-1}(j-1) q_{n-1}\right) \bmod q_{n, i}, j=1,2, \ldots, q_{n, i} \tag{13}
\end{equation*}
$$

For any other value of $N$, the circle is composed of gaps of three different lengths.

### 3.2 The String of Gap Types

Now let us consider the more dynamic situation-we will describe the change in gap structure induced by the addition of extra points. In particular, we are interested in the transition from a circle of $q_{n-1}$ gaps to one of $q_{n}$ gaps. Notation is needed.

Suppose the circle is partitioned into gaps of only two different lengths, say large and small. We label a large gap $l$ and a small gap $s$. Let

$$
\Phi_{n}=\phi_{n, 1} \phi_{n, 2} \cdots \phi_{n, q_{n}}
$$

denote the string of gap types when $N=q_{n}$, ordered clockwise from the origin
around the circle so that $\phi_{n, j}$ denotes the gap type (either $s$ or $\eta$ ) of the $j$ th gap formed by points $u_{j}$ and $u_{j+1}$. Assume that $\Phi_{0}=s$.

For any string $S$ and nonnegative integer $t$, denote by $S^{t}$ the concatenation of $S$ with itself $t$ times, where $S^{0}$ is the empty string. For any strings $S_{1}, S_{2}$, we write $S_{1} S_{2}$ for the concatenation of $S_{1}$ followed by $S_{2}$. Define $P_{n}$ such that

$$
P_{n}(\eta)=\left\{\begin{array}{ll}
s^{a_{n}} \tau, & n \text { odd }, \\
Z s^{a_{n}}, & n \text { even },
\end{array} \quad P_{n}(s)= \begin{cases}s^{a_{n}-1} \tau, & n \text { odd }, \\
\tau s^{a_{n}-1}, & n \text { even } .\end{cases}\right.
$$

The following theorem shows that $P_{n}$ is the production rule which describes the manner in which the string of gap types develops as more points are included on the circle. The result may, after a little effort, be derived from (12). We omit the proof, and refer the reader to Theorem 4.1 in [18].

Theorem 1:

$$
\Phi_{n}=P_{n}\left(\Phi_{n-1}\right)=P_{n}\left(\phi_{n-1,1}\right) P_{n}\left(\phi_{n-1}, 2\right) \ldots P_{n}\left(\phi_{n-1}, q_{n-1}\right) .
$$

Example: For the golden number, $\tau=(\sqrt{5}-1) / 2$,

$$
P_{n}(\tau)=\left\{\begin{array}{ll}
s l, & n \text { odd }, \\
l_{s}, & n \text { even },
\end{array} \quad P_{n}(s)=\tau\right.
$$

Hence,

$$
\begin{aligned}
& \Phi_{0}=s, \\
& \Phi_{1}=l, \\
& \Phi_{2}=l s, \\
& \Phi_{3}=s l l, \\
& \Phi_{4}=l z s l s, \\
& \Phi_{5}=s l s l z s l l .
\end{aligned}
$$

We now introduce the following two results which we will need to prove our main result in Section 4.2. Proposition 2 demonstrates a simple property of the production rule $P_{n}$, while Proposition 3 shows that a component of the string $\Phi_{n}$ is symmetric.

Let $\theta=\theta_{1} \theta_{2} \ldots \theta_{k}$ denote a string of $k$ letters, where $\theta_{i}=s$ or $l, i=1$, $2, \ldots, k$. For any string $S$, let $S^{*}$ denote the string $S$ in reverse order. We write $P_{n}(\theta)^{*}$ and $P_{n}\left(\theta_{k}\right)^{*}$ for $\left(P_{n}(\theta)\right)^{*}$ and $\left(P_{n}\left(\theta_{k}\right)\right)^{*}$, respectively.
Proposition 2: $P_{n}(\theta)^{*}=P_{n-1}\left(\theta^{*}\right)$.

$$
\text { Proof: } \quad \begin{aligned}
P_{n}(\theta) * & =\left(P_{n}\left(\theta_{1}\right) P_{n}\left(\theta_{2}\right) \ldots P_{n}\left(\theta_{k}\right)\right)^{*}, \\
& =P_{n}\left(\theta_{k}\right){ }^{*} P_{n}\left(\theta_{k-1}\right) * \ldots P_{n}\left(\theta_{1}\right)^{*}, \\
& =P_{n-1}\left(\theta_{k}\right) P_{n-1}\left(\theta_{k-1}\right) \ldots P_{n-1}\left(\theta_{1}\right), \\
& =P_{n-1}\left(\theta_{k} \theta_{k-1} \cdots \theta_{1}\right), \\
& =P_{n-1}\left(\theta^{*}\right),
\end{aligned}
$$

where we have used the fact that $P_{n}(s)^{*}=P_{n-1}(s)$ and $P_{n}(\Omega)^{*}=P_{n-1}(l)$.

$$
\text { Let } B_{n}=\phi_{n, 2} \phi_{n, 3} \cdots \phi_{n}, q_{n}-1 .
$$

Proposition 3: $B_{n}^{*}=B_{n}(n \geq 1)$.
Proof: For $n=1$, the result is trivial since from Theorem $1, \Phi_{1}=s^{a_{1}-1} l$. Now consider the case $n \geq 2$. It is necessary to show that

$$
\phi_{n, j}=\phi_{n, q_{n}-j+1}, \quad j=2,3, \ldots, q_{n}-1
$$

From (13), with $i=a_{n}$ (since $N=q_{n}$ ),

$$
\begin{equation*}
u_{j}=\left((-1)^{n-1}(j-1) q_{n-1}\right) \bmod q_{n}, j=1,2, \ldots, q_{n} . \tag{14}
\end{equation*}
$$

Thus,

$$
\begin{aligned}
u_{q_{n}-j+2} & =\left((-1)^{n-1}\left(q_{n}-(j-1)\right) q_{n-1}\right) \bmod q_{n}, \\
& =\left((-1)^{n}(j-1) q_{n-1}\right) \bmod q_{n} \quad\left(j=2,3, \ldots, q_{n}\right) .
\end{aligned}
$$

Hence，

$$
u_{j}+u_{q_{n}-j+2}=q_{n} \quad\left(j=2,3, \ldots, q_{n}\right)
$$

Thus，

$$
\begin{aligned}
u_{j+1}-u_{j} & =q_{n}-u_{q_{n}-j+1}-\left(q_{n}-u_{q_{n}-j+2}\right)\left(j=2,3, \ldots, q_{n}-1\right), \\
& =u_{q_{n}-j+2}-u_{q_{n}-j+1},
\end{aligned}
$$

from which the result follows．

## 4．The Characteristic of $\alpha$

## 4．1 General $\alpha$

The following method of constructing the characteristic is described in Venkov（［20］，pp．65－68）．Markoff first showed that the characteristic of $\alpha$ is equal to $\beta_{1} \beta_{2} \beta_{3} \ldots$ ，where

$$
\beta_{n}=\beta_{n-1}^{a_{n}-1} \beta_{n-2} \beta_{n-1}, \beta_{0}=s, \beta_{1}=s^{\alpha_{1}-1} \eta .
$$

We mention that if $\alpha$ is rational，say $\alpha=\left\{\alpha_{0} ; \alpha_{1}, \alpha_{2}, \ldots, \alpha_{N}\right\}$ ，then $\beta_{1} \beta_{2}$ $\ldots \beta_{N-1}\left(\beta_{N}\right)^{\infty}$ is the characteristic where $N$ is even（so that the number of terms is odd）．If $N$ is odd，the number of terms can be made odd，as ．．．$\alpha_{N-1}$ ， $\left.\alpha_{N}\right\}$ can be replaced by $\left.\ldots \alpha_{N-1}, \alpha_{N}-1,1\right\}$ ，if $\alpha_{N}>1$ ．If $\alpha_{N}=1$（and $\alpha \neq 1$ ）， then $\left.\ldots \alpha_{N-2}, \alpha_{N-1}, \alpha_{N}\right\}$ can be replaced by $\left.\ldots \alpha_{N-2}, \alpha_{N-1}+1\right\}$ ．

Let $\alpha=\{0 ; 1,2,3\}=\{0 ; 1,2,2,1\}=7 / 10$ ．Then
$\beta_{0}=s$ ，
$\beta_{1}=\tau$ ，
$\beta_{2}=\beta_{1} \beta_{0} \beta_{1}=l s l$,
$\beta_{3}=\beta_{2} \beta_{1} \beta_{2}=$ lsillss ，

The characteristic is then given by $\beta_{1} \beta_{2} \beta_{3}\left(\beta_{4}\right)^{\infty}$ ，that is，
こてsてZsてZてsて(ZsてZsてZてsて)

Fraenkel et al．（［4］，Theorem 1）offer an alternative method of construc－ tion：they show that the characteristic is equal to $\lim _{n \rightarrow \infty} \delta_{n}$ ，where

$$
\begin{equation*}
\delta_{n}=\delta_{n-1}^{a_{n}} \delta_{n-2}, \delta_{0}=s, \delta_{1}=s^{a_{1}-1} \eta . \tag{16}
\end{equation*}
$$

They actually form the characteristic by means of＂shift operators．＂It may be shown，however，that the recurrence relation（16）is an equivalent means of formulating the characteristic，in terms of the actual operations required．

Note that if $\alpha=\left\{\alpha_{0} ; \alpha_{1}, \alpha_{2}, \ldots, \alpha_{N}\right\}$ ，then $\delta_{N}^{\infty}$ is the characteristic（if $N$ is even）．
Example：As in the above example，consider $\alpha=7 / 10$ ．Then

$$
\begin{aligned}
& \delta_{0}=s, \\
& \delta_{1}=2, \\
& \delta_{2}=2 Z s, \\
& \delta_{3}=2 Z s i z s l, \\
& \delta_{4}=2 \tau s Z 2 s i z s .
\end{aligned}
$$

Thus，$\delta_{4}^{\infty}$ is the characteristic．
The method of Fraenkel et al．［4］generalizes the work done by Stolarsky （［12］，Theorem 2），who shows how to generate the characteristic for the parti－ cular case where $\alpha=\{1 ; a, \alpha, \ldots\}$ ，the positive root of $x^{2}+(\alpha-2) x-\alpha=0$ ．

## CHARACTERISTICS AND THE THREE GAP THEOREM

In this paper, we present a new proof of Theorem 1 in [4] and Theorem 2 in [12] for the case $\alpha=\{0, \alpha, \alpha, \ldots\}$, the positive root of $x^{2}+\alpha x-1=0$ by exhibiting a connection between the characteristic of $\alpha$ and its string of gap types (see Theorems 5 and 8 below).

### 4.2 The Characteristic of $\alpha=\{0 ; a, a, \ldots\}$

From now on, unless otherwise stated, assume that $\alpha=\{0 ; \alpha, \alpha, \ldots\}$. For this case, we show how the string of gap types $\Phi_{n}$ is generated recursively and how it is related to the characteristic of $\alpha$.
Theorem 4:

$$
\Phi_{n}=\Phi_{n-1}^{* a} \Phi_{n-2}(n \geq 2), \Phi_{0}=s, \Phi_{1}=s^{a-1} l
$$

Proof: Theorem 1 implies the truth of the assertion for $n=2$, 3 . Using the induction hypothesis we show that the result holds in general by verifying it for $n=k+1$, assuming that it holds for $n=k$ and $n=k-1$.

$$
\begin{aligned}
\Phi_{k+1} & =P_{k+1}\left(\Phi_{k}\right) \\
& =P_{k+1}\left(\Phi_{k-1}^{* a} \Phi_{k-2}\right) \\
& =P_{k+1}\left(\Phi_{k-1}^{* a}\right) P_{k+1}\left(\Phi_{k-2}\right) \\
& =P_{k-1}\left(\Phi_{k-1}^{*}\right)^{a} P_{k-1}\left(\Phi_{k-2}\right)
\end{aligned}
$$

Thus,

$$
\Phi_{k+1}=\Phi_{k}^{*} \alpha_{\Phi_{k-1}}
$$

which follows from Theorem 1 and Proposition 2 for $\theta=\Phi_{k-1}$ and the fact that $P_{k+1}=P_{k-1}$ for all $k \geq 2$.

The following theorem shows how the string $\Phi_{n}$ is related to another string $\Omega_{n}$ which corresponds to the first $q_{n}$ elements of the characteristic. One merely places the first element of $\Phi_{n}$ in the penultimate position of $\Phi_{n}$ to obtain $\Omega_{n}$.

We let $A_{n}=\phi_{n, 1}, B_{n}$ be as in Proposition 3, and $C_{n}=\phi_{n, q_{n}}$. Now, let

$$
\Omega_{n}=B_{n} A_{n} C_{n}
$$

Theorem 5:

$$
\Omega_{n}=\Omega_{n-1}^{a} \Omega_{n-2}(n \geq 2), \Omega_{0}=s, \Omega_{1}=s^{\alpha-1} l^{\prime}
$$

Proof: The result is readily shown to be true for $n=2$, 3 from direct observation of the strings $\Omega_{2}$ and $\Omega_{3}$. These strings derive from $\Phi_{2}$ and $\Phi_{3}$, which may be written down using Theorem 4. In what follows, assume that $n>3$.

Induction on $n$ using Theorem 1 implies

$$
A_{n}=C_{n-1}= \begin{cases}s, & n \text { odd }  \tag{17}\\ \tau, & n \text { even }\end{cases}
$$

(This is actually true for $n \geq 2$.)
We are required to show that

$$
B_{n} A_{n} C_{n}=\left(B_{n-1} A_{n-1} C_{n-1}\right)^{a} B_{n-2} A_{n-2} C_{n-2},
$$

or, using (17),

$$
\begin{equation*}
B_{n}=\left(B_{n-1} A_{n-1} C_{n-1}\right)^{a} B_{n-2} \tag{18}
\end{equation*}
$$

Theorem 4 is equivalent to the statement

$$
A_{n} B_{n} C_{n}=\left(C_{n-1} B_{n-1}^{*} A_{n-1}\right)^{a} A_{n-2} B_{n-2} C_{n-2}
$$

Using (17) and rearranging terms leads to

$$
B_{n}=B_{n-1}^{*} A_{n-1}\left(C_{n-1} B_{n-1}^{*} A_{n-1}\right)^{a-1} A_{n-2} B_{n-2}
$$

Further manipulation gives

$$
B_{n}=\left(B_{n-1}^{*} A_{n-1} C_{n-1}\right)^{a} B_{n-2}
$$

which is equivalent to (18). (Recall from Proposition 3 that $B_{n}^{*}=B_{n}$.) Thus, the theorem is proved.

The following corollary gives the production rule for the string $\Omega_{0} \Omega_{1} \Omega_{2} \ldots$ The proof is by induction and is omitted. Note that the production rule is independent of $n$.
Corollary 6: Suppose that $Q(s)=s^{\alpha-1} \eta, Q(\eta)=s^{\alpha-1} \mathcal{l}$. Then

$$
\Omega_{n}=Q\left(\Omega_{n-1}\right)=\Omega_{n-1}^{a} \Omega_{n-2}(n \geq 2), \Omega_{0}=s, \Omega_{1}=s^{a-1} \downarrow
$$

Example: For $\alpha=\tau$, we have $Q(s)=\tau, Q(\tau)=\tau_{s}$, and $\Omega_{n}=\Omega_{n-1} \Omega_{n-2}$ for $n \geq 1$. Hence,

$$
\begin{aligned}
& \Omega_{0}=s, \\
& \Omega_{1}=l, \\
& \Omega_{2}=l_{s}, \\
& \Omega_{3}=l_{s} l \\
& \Omega_{4}=l_{s} l Z_{s}, \\
& \Omega_{5}=l_{s} l l_{s} l_{s} l_{0} .
\end{aligned}
$$

The Golden Sequence is then $\lim _{n \rightarrow \infty} \Omega_{n}$. Comparing Theorem 5 with Fraenkel et al's result ([4], Theorem 1) [equivalent to our Equation (16)] identifies $\Omega_{n}$ as the first $q_{n}$ elements of the characteristic. That is, $\Omega_{n}=\delta_{n}$, where $\delta_{n}$ is defined by (16). Thus, the string of gap types is generated in the same way as the characteristic, a result all the more surprising since it does not hold for all $\alpha$. We proceed to verify the connection between $\Omega_{n}$ and the characteristic by exploiting the relationship between $\Phi_{n}$ and $\Omega_{n}$. This, then (with Theorem 5), forms the new proof of Theorem 1 in [4] and Theorem 2 in [12] for the case $\alpha=$ $\{0 ; a, a, \ldots\}$. The proof sheds light on the set of numbers for which the string of gap types corresponds to the characteristic. First, we need the following, which is proved in van Ravenstein ([17], Equation 5.12).
Lemma 7: $[k \alpha]=\left[k \frac{p_{n, i}}{q_{n, i}}\right], k=1,2, \ldots, q_{n, i}-1\left(n \geq 2,1 \leq i \leq a_{n}\right)$, where $\alpha$ is any irrational number.

$$
\text { Let } \Omega_{n}=\omega_{n, 1} \omega_{n}, 2 \ldots \omega_{n}, q_{n} .
$$

Theorem 8: For $n \geq 2$,

$$
\omega_{n, j}= \begin{cases}s, & d_{j}=0 \\ l, & d_{j}=1\end{cases}
$$

where $d_{j}$ is defined by (1) and $j=1,2, \ldots, q_{n}$.
Proof: Equation (14) is equivalent to

$$
\begin{aligned}
u_{j} & =q_{n}\left\{(-1)^{n-1}(j-1) \frac{q_{n-1}}{q_{n}}\right\} \\
& =(-1)^{n-1}(j-1) q_{n-1}-q_{n}\left[(-1)^{n-1}(j-1) \frac{q_{n-1}}{q_{n}}\right] .
\end{aligned}
$$

Thus, for $j=2,3, \ldots, q_{n}-1$,

$$
\begin{align*}
u_{j+1}-u_{j} & =(-1)^{n-1} q_{n-1}-(-1)^{n-1} q_{n}\left(\left[j \frac{q_{n-1}}{q_{n}}\right]-\left[(j-1) \frac{q_{n-1}}{q_{n}}\right]\right),  \tag{19}\\
& =(-1)^{n-1} q_{n-1}-(-1)^{n-1} q_{n}([j \alpha]-[(j-1) \alpha]) \tag{20}
\end{align*}
$$

The latter step follows from Lemma 7 and Equation (11). Hence, for $j=2,3$, $\ldots, q_{n}-1$,

$$
u_{j+1}-u_{j}= \begin{cases}(-1)^{n-1} q_{n-1}, & d_{j-1}=0 \\ (-1)^{n-1}\left(q_{n-1}-q_{n}\right), & d_{j-1}=1\end{cases}
$$

From (9) and (10), it may now be shown that

$$
\phi_{n, j}= \begin{cases}s, & d_{j-1}=0  \tag{21}\\ l, & d_{j-1}=1\end{cases}
$$

where $j=2,3, \ldots, q_{n}-1$.
To complete the proof, first note from (6) that

$$
\left[q_{n} \alpha\right]= \begin{cases}p_{n}-1, & n \text { odd } \\ p_{n}, & n \text { even }\end{cases}
$$

From Lemma 7, $\left[\left(q_{n}-1\right) \alpha\right]=\left[\left(q_{n}-1\right) p_{n} / q_{n}\right]=p_{n}-1$. Therefore,

$$
d_{q_{n}-1}=\left[q_{n} \alpha\right]-\left[\left(q_{n}-1\right) \alpha\right]= \begin{cases}0, & n \text { odd } \\ 1, & n \text { even }\end{cases}
$$

From (14) and (9) we have

$$
\phi_{n, 1}= \begin{cases}s, & d_{q_{n}-1}=0  \tag{22}\\ z, & d_{q_{n}-1}=0\end{cases}
$$

The result for $\phi_{n, q_{n}}$ follows similarly. From Lemma A1 (see Appendix), $\left[\left(q_{n}+1\right) \alpha\right]=p_{n} \cdot$ Hence,

$$
d_{q_{n}}=\left[\left(q_{n}+1\right) \alpha\right]-\left[q_{n} \alpha\right]= \begin{cases}0, & n \text { even } \\ 1, & n \text { odd }\end{cases}
$$

and thus, from (14) and (9),

$$
\phi_{n, q_{n}}= \begin{cases}s, & d_{q_{n}}=0  \tag{23}\\ \tau, & d_{q_{n}}=1\end{cases}
$$

Theorem 5 and Equations (21)-(23) establish the proof.
Corollary 9: Suppose that $\alpha=\left\{0 ; a_{1}, \alpha_{2}, \ldots\right\}$, where $a_{j}=\alpha_{i-j+1}$ for $j=1,2$, ..., i. Then

$$
\omega_{i, j}= \begin{cases}s, & d_{j}=0 \\ z, & d_{j}=1\end{cases}
$$

Proof: For this value of $\alpha$,

$$
\frac{p_{i}}{q_{i}}=\frac{q_{i-1}}{q_{i}}=\left\{0 ; a_{i}, a_{i-1}, \ldots, a_{2}, a_{1}\right\}
$$

The proof is then identical to the proof of Theorem 8; in particular, the step from (19) to (20) follows. $\square$

The correspondence between $\Phi_{n}$ and the characteristic does not hold for all $\alpha$, as the following (counter)example shows.

$$
\text { Let } \alpha=\{0 ; 1,2,3,1+\tau\}=\frac{2 \tau+9}{3 \tau+13} \text {. Then }
$$

$$
\Phi_{0}=s, \Phi_{1}=l, \Phi_{2}=l s s, \Phi_{3}=s s s l_{s s} l_{s s} l,
$$

and thus,

$$
\Omega_{0}=s, \Omega_{1}=l, \Omega_{2}=s l s, \Omega_{3}=s s l_{s s} l_{s s s} l,
$$

which does not correspond to the characteristic, since

$$
\delta_{0}=s, \delta_{1}=l, \delta_{2}=l l s, \delta_{3}=\eta l s l l s l l s \tau .
$$

Conjecture: The correspondence between $\Phi_{n}$ and the characteristic holds only for $\alpha$ equivalent to the number $\{0 ; \alpha, \alpha, \alpha, \ldots\}$.

APPENDIX. The Evaluation of $[N \alpha], N=1,2, \ldots$
We have shown how one may evaluate the integer parts of positive consecutive multiples of a number by forming its characteristic. Here, we present an alternative method by which we decompose the number into terms related to its continued fraction expansion. The method appears in Fraenkel et al. [3] and is central to their paper. We offer a new and shorter proof.

Lemma A1 (see Fraenkel et a.1.[4], Lemma 2): Suppose that $n>0$ and $0<q<q_{n}$. Then $\left[\left(q+q_{n-1}\right) \alpha\right]=p_{n-1}+\left[q^{\alpha}\right]$.
Lemma A2 (see, e.g., Fraenkel [5], Theorem 3): There is a unique decomposition of any natural number $N$ in the form

$$
N=\sum_{i=0}^{m} b_{i} q_{i},
$$

where the $b_{i}^{\prime}$ s are integers; $0 \leq b_{0}<q_{1}, 0 \leq b_{i} \leq a_{i+1}, i>0$, and $b_{i}=a_{i+1}$, only if $b_{i-1}=0$. Since this expansion is unique,

$$
\begin{equation*}
\sum_{i=0}^{n} b_{i} q_{i}<q_{n+1} \tag{A}
\end{equation*}
$$

Theorem A3: If $N=\sum_{i=k}^{m} b_{i} q_{i}$, then

$$
[N \alpha]= \begin{cases}\sum_{i=k}^{m} b_{i} p_{i}, & k \text { even }, \\ -1+\sum_{i=k}^{m} b_{i} p_{i}, & k \text { odd }\end{cases}
$$

where $b_{k} \neq 0$ (i.e., $k=\max \left\{j: b_{j}>0\right\}$ ).
Proof: If $N=\sum_{i=k}^{m} b_{i} q_{i}, b_{k} \neq 0$, then

$$
[N \alpha]=\left[\left(\sum_{i=k}^{m-1} b_{i} q_{i}+b_{m} q_{m}\right) \alpha\right]=\left[\left(\sum_{i=k}^{m-1} b_{i} q_{i}+\left(b_{m}-1\right) q_{m}+q_{m}\right) \alpha\right]
$$

From (A),

$$
\sum_{i=k}^{m-1} b_{i} q_{i}+\left(b_{m}-1\right) q_{m}<b_{m} q_{m} \leq a_{m+1} q_{m}<q_{m+1} .
$$

Hence, from Lemma Al,

$$
[N \alpha]=p_{m}+\left[\left(\sum_{i=k}^{m-1} b_{i} q_{i}+\left(b_{m}-1\right) q_{m}\right) \alpha\right]
$$

Further application of (A) and the lemma leads to

$$
[N \alpha]=b_{m} p_{m}+\left[\sum_{i=k}^{m-1} b_{i} q_{i} \alpha\right]
$$

Clearly, we are led to

$$
[N \alpha]=\sum_{i=k+1}^{m} b_{m} p_{m}+\left[b_{k} q_{k} \alpha\right]
$$

From (6),

$$
b_{k} q_{k} \alpha-b_{k} p_{k}=\frac{b_{k}(-1)^{k}}{t_{k+1} p_{k}+p_{k+1}}
$$

Thus, $-1<b_{k}\left(q_{k} \alpha-p_{k}\right)<1$, since $0 \leq b_{k} \leq a_{k+1}$. Hence,

$$
\left[b_{k} q_{k} \alpha\right]= \begin{cases}b_{k} p_{k}, & k \text { even } \\ b_{k} p_{k}-1, & k \text { odd }\end{cases}
$$

This completes the proof. $\square$

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## Announcement

## FOURTH INTERNATIONAL CONFERENCE ON FIBONACCI NUMBERS AND THEIR APPLICATIONS

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# A TWO-VARIABLE LAGRANGE-TYPE INVERSION FORMULA WITH APPLICATIONS TO EXPANSION AND CONVOLUTION IDENTITIES 

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## 1. Introduction

Using Lagrange inversion, one obtains the formal power series (fps) expansions (cf. Riordan [12], Sec. 4.5)

$$
\begin{align*}
& \exp b z=\sum_{k=0}^{\infty} \frac{b(a k+b)^{k-1}}{k!} w^{k}  \tag{1.1}\\
& \frac{\exp b z}{1-\alpha z}=\sum_{k=0}^{\infty} \frac{(a k+b)^{k}}{k!} w^{k} \tag{1.2}
\end{align*}
$$

where $w=z \cdot \exp (-\alpha z)$ s and

$$
\begin{align*}
& (1+z)^{b}=\sum_{k=0}^{\infty} \frac{b}{a k+b}\binom{a k+b}{k} v^{k}  \tag{1.3}\\
& \frac{(1+z)^{b}}{1-\frac{a z}{1+z}}=\sum_{k=0}^{\infty}\binom{a k+b}{k} v^{k}
\end{align*}
$$

where $v=z /(1+z)^{a}$. With the help of these identities, Gould [6-8] obtained many convolution identities. Higher-dimensional extensions of (1.2) and (1.4) were studied and proved by Carlitz [1, 2] using MacMahon ${ }^{\text {s }}$ Master Theorem. Finally, Carlitz's identities were embedded into a general theory by Joni [9]. The key for her results, again, is Lagrange inversion (this time the multivariable Lagrange-Good inversion formula, cf. Joni [10]).

In [5] Cohen \& Hudson discovered two-variable generalizations of (1.1) and (1.2) that are different in nature from the corresponding results of Carlitz, and studied related convolution identities. Their proofs are based on a specific operator method also used in Cohen's papers [3] and [4]. Thus, the question remained open as to whether there might be a Lagrange-type inversion formula providing the background for Cohen \& Hudson's results and yielding twovariable extensions of (1.3) and (1.4), in addition. This formula will be given in Section 3 (Theorem 1, Corollary 2) of this paper. Subsequently, we are able to derive all of Cohen \& Hudson's results and, moreover, to give the "factorial" analogues that correspond to (1.3) and (1.4). This will be done in Section 4. For the purpose of illustration, we list some identities in the next section.

## 2. Some Expansion and Convolution Identities

To write our identities, it is convenient to adopt the usual multidimensional notations. Let $\underset{\sim}{\mathcal{K}}=\left(k_{1}, \mathcal{K}_{2}\right), \underset{\sim}{n}=\left(n_{1}, n_{2}\right) \in \mathbb{Z}^{2}$ (pairs of integers) and $\underset{\sim}{\underset{\sim}{A}}=\left(z_{1}, z_{2}\right)$ be a pair off indeterminat̃es, then we define $\underset{\sim}{k}!=k_{1}!k_{2}!, \underset{\sim}{n} \geq \underset{\sim}{k}$ if and only if $n_{1} \geq k_{1}$ and $n_{2} \geq k_{2}, \underset{\sim}{n}-\underset{\sim}{k}=\left(n_{1}-k_{1}, n_{2}-k_{2}^{\sim}\right), \underset{\sim}{0}=(0,0)$,

$$
\underset{\sim}{\underset{\sim}{z}}=z_{1}^{k_{1}} z_{2}^{k_{2}} \quad \text { and } \quad(\underset{\sim}{\underset{\sim}{\tilde{K}}})=\binom{z_{1}}{k_{1}}\binom{z_{2}}{k_{2}} .
$$

Throughout this paper, for $\underset{\sim}{k} \in \mathbb{Z}^{2}, \lambda_{i}, \mu_{i}, \alpha_{i}, \beta_{i} \in \mathbb{C}$ (complex numbers), $i=1$, 2, we shall write

$$
\begin{array}{ll}
R_{1}(\underset{\sim}{k})=\frac{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\alpha_{2}+\beta_{2} k_{2}\right)}{\lambda_{2}+\mu_{2} k_{2}}, & R_{2}(\underset{\sim}{k})=\frac{\left(\lambda_{2}+\mu_{2} k_{2}\right)\left(\alpha_{1}+\beta_{1} k_{1}\right)}{\lambda_{1}+\mu_{1} k_{1}}, \\
r_{1}(\underset{\sim}{k})=\frac{\left(\lambda_{1}+\mu_{1} k_{1}\right)}{\left(\lambda_{2}+\mu_{2} k_{2}\right)}, & r_{2}(\underset{\sim}{k})=\frac{\left(\lambda_{2}+\mu_{2} k_{2}\right)}{\left(\lambda_{1}+\mu_{1} k_{1}\right)},
\end{array}
$$

for short. Note that $r_{i}(\underset{\sim}{k})$ is equal to $R_{i}(\underset{\sim}{k})$ with $\beta_{i}=0$ and $\alpha_{i}=1$.
The first of Cohen \& Hudson's identities, (1.3) in [5], is equivalent to

$$
\begin{equation*}
\frac{1-\frac{\mu_{1} \mu_{2}}{\lambda_{1} \lambda_{2}} z_{1} z_{2}}{\left(1-\frac{\mu_{1}}{\lambda_{2}} z_{1}\right)\left(1-\frac{\mu_{2}}{\lambda_{1}} z_{2}\right)}=\sum_{\underset{\sim}{k} \geq 0} \frac{r_{1}(\underset{\sim}{k})^{k_{1}}}{k_{1}!} \frac{r_{2}(\underset{\sim}{k})^{k_{2}}}{k_{2}!}{\underset{\sim}{c}}_{\underset{\sim}{k}}^{k} \exp \left(-r_{1}(\underset{\sim}{k}) z_{1}-r_{2}(\underset{\sim}{k}) z_{2}\right) . \tag{2.1}
\end{equation*}
$$

For $z_{2}=0$, (2.1) reduces to (1.2). The factorial analogue of (2.1) we prove is
(2.2) $\frac{1-\frac{\mu_{1} \mu_{2}}{\lambda_{1} \lambda_{2}} \frac{z_{1} z_{2}}{\left(1+z_{1}\right)\left(1+z_{2}\right)}}{\left(1-\frac{\mu_{1}}{\lambda_{2}} \frac{z_{1}}{1+z_{1}}\right)\left(1-\frac{\mu_{2}}{\lambda_{1}} \frac{z_{2}}{1+z_{2}}\right)}$

$$
=\sum_{\underset{\sim}{k} \geq 0}\binom{r_{1}(\underset{\sim}{k}}{k_{1}^{k}}\binom{r_{2}(\underset{\sim}{k})}{k_{2}} \underset{\sim}{z}\left(1+z_{1}\right)^{-r_{1}(\underset{\sim}{k})}\left(1+z_{2}\right)^{-r_{2}(\underset{\sim}{k})} .
$$

Equation (2.2) reduces to (1.4) for $z_{2}=0$. The "mixed" expansion,

$$
\begin{equation*}
\frac{1-\frac{\mu_{1} \mu_{2}}{\lambda_{1} \lambda_{2}} z_{1} \frac{z_{2}}{1+z_{2}}}{\left(1-\frac{\mu_{1}}{\lambda_{2}} z_{1}\right)\left(1-\frac{\mu_{2}}{\lambda_{1}} \frac{z_{2}}{1+z_{2}}\right)}+\sum_{\underset{\sim}{k} \geq 0} \frac{r_{1}(\underset{\sim}{k})^{k_{1}}}{k_{1}!}\binom{r_{2}(\underset{\sim}{k})}{k_{2}^{k}} z_{\sim}^{k} \exp \left(-r_{1}(\underset{\sim}{k})\right)\left(1+z_{2}\right)^{-r_{2}(\underset{\sim}{k})}, \tag{2.3}
\end{equation*}
$$

is a two-variable generalization of (1.2) and (1.4) at the same time. This is seen by setting $z_{2}=0$ or $z_{1}=0$, respectively.

The second expansion of Cohen $\&$ Hudson, (1.5) in [5], is equivalent to
(2.4) $\frac{1}{\lambda_{1}-\mu_{2} \alpha_{1} z_{2}}=\sum_{\underset{\sim}{k} \geq \underset{\sim}{0}} \frac{1}{\lambda_{1}+\mu_{1} k_{1}} \frac{r_{1}(\underset{\sim}{k})^{k_{1}} R_{2}(\underset{\sim}{k})^{k_{2}}}{\underset{\sim}{k}!} \underset{\sim}{z} \underset{\sim}{k} \exp \left(-x_{1}(\underset{\sim}{k}) z_{1}-R_{2}(\underset{\sim}{k}) z_{2}\right)$.

Setting $z_{2}=0$ in (2.4) gives (1.1); setting $z_{1}=0$ gives (1.2). The factorial analogue of (2.4) is presented here:
(2.5) $\frac{1}{\lambda_{1}-\mu_{2} \alpha_{1} \frac{z_{2}}{1+z_{2}}}=\sum_{\underset{\sim}{k} \geq 0} \frac{1}{\lambda_{1}+\mu_{1} k_{1}}\binom{r_{1}}{{\underset{\sim}{k}}_{1}^{(k)}}\left(\underset{\underset{\sim}{k_{2}}}{R_{2}(\underset{\sim}{k})}\right)_{\sim}^{z} \underset{\sim}{k}\left(1+z_{1}\right)^{-r_{1}(\underset{\sim}{k})}\left(1+z_{2}\right)^{-R_{2}(\underset{\sim}{\sim})}$. This is a generalization of (1.3) and (1.4) at the same time. Similarly, a two-dimensional generalization of (1.1) and (1.4) is

$$
\begin{align*}
& \frac{1}{\lambda_{1}-\mu_{2} \alpha_{1} \frac{z_{2}}{1+z_{2}}}  \tag{2.6}\\
& =\sum_{\underset{\sim}{k} \geq{\underset{\sim}{0}} \frac{1}{\lambda_{1}+\mu_{1} k_{1}} \frac{r_{1}(\underset{\sim}{k})^{k_{1}}}{k_{1}!}\left({\underset{\sim}{k}}_{2}^{R_{2}(\underset{\sim}{k})}\right) \underset{\sim}{z} \underset{\sim}{k} \exp \left(-r_{1}(\underset{\sim}{k}) z_{1}\right)\left(1+z_{2}\right)^{-R_{2}(\underset{\sim}{k})},} .
\end{align*}
$$

and a generalization of (1.2) and (1.3) is
(2.7) $\frac{1}{\lambda_{1}-\mu_{2} \alpha_{1} z_{2}}=\sum_{\underset{\sim}{k} \geq{\underset{\sim}{0}}_{0}} \frac{1}{\lambda_{1}+\mu_{1} k_{1}}\left({\underset{\sim}{1}}_{r_{1}(\underset{\sim}{k})}^{k_{1}}\right) \frac{R_{2}(\underset{\sim}{k})^{k_{2}}}{k_{2}!} \underset{\sim}{z}\left(1+z_{1}\right)^{-r_{1}(\underset{\sim}{k})} \exp \left(-R_{2}(\underset{\sim}{k}) z_{2}\right)$.

We give another expansion, which Cohen \& Hudson missed:

$$
\begin{equation*}
1=\sum_{\underset{\sim}{k} \geq 0} \frac{\lambda_{1} \lambda_{2}-k_{1} k_{2} \frac{\left(\alpha_{1} \mu_{1}-\beta_{1} \lambda_{1}\right)\left(\alpha_{2} \mu_{2}-\beta_{2} \lambda_{2}\right)}{\left(\alpha_{1}+\beta_{1} k_{1}\right)\left(\alpha_{2}+\beta_{2} k_{2}\right)}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)} \frac{R_{1}(\underset{\sim}{k})^{k_{1}} R_{2}(\underset{\sim}{k})^{k_{2}}}{\underset{\sim}{k}!} \cdot \tag{2.8}
\end{equation*}
$$

This is a two-variable generalization of (1.1). The corresponding generalization of (1.3) reads

$$
\begin{equation*}
I=\sum_{\underset{\sim}{k} \geq 0} \frac{\lambda_{1} \lambda_{2}-k_{1} k_{2} \frac{\left(\alpha_{1} \mu_{1}-\beta_{1} \lambda_{1}\right)\left(\alpha_{2} \mu_{2}-\beta_{2} \lambda_{2}\right)}{\left(\alpha_{1}+\beta_{1} k_{1}\right)\left(\alpha_{2}+\beta_{2} k_{2}\right)}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)}\binom{R_{1}(\underset{\sim}{k})}{k_{1}}\binom{R_{2}(\underset{\sim}{k})}{k_{2}} . \tag{2.9}
\end{equation*}
$$

that of (1.1) and (1.3),

$$
\cdot \underset{\sim}{\underset{\sim}{z}}\left(1+z_{1}\right)^{-R_{1}(\underset{\sim}{k})}\left(1+z_{2}\right)^{-R_{2}(\underset{\sim}{k})},
$$



Identities (2.1)-(2.10) will be proved in Section 4 by establishing three types of general expansions that underly (2.1)-(2.3), (2.4)-(2.7), and (2.8)-(2.10), respectively.

From each of the expansions (2.1)-(2.10), we may derive a convolution identity which generalizes Jensen's convolution or the Abel-type Jensen-Gould convolution identity, respectively (cf. [6], [7]). We shall give two examples; the remaining identities are obtained similarly.

Multiplying both sides of (2.1) with $\exp \left(s_{1} z_{1}+s_{2} z_{2}\right)$ and comparing coefficients of $\underset{\sim}{\underset{n}{n}}$, we obtain (1.4) of [5].

The factorial analogue of (2.11), deduced by multiplying both sides of (2.2) with $\left(1+z_{1}\right)^{s_{1}}\left(1+z_{2}\right)^{s_{2}}$ and comparing coefficients of $\underset{\sim}{z}$ reads

## 3. Lagrange Inversion

Let

$$
\phi(z)=\left(\phi_{1}\left(z_{1}, z_{2}\right), \phi_{2}\left(z_{1}, z_{2}\right)\right)
$$

be some pair of $f p s$ in $z_{1}$ and $z_{2}$ with $\phi_{i}(0,0) \neq 0, i=1,2$. Let

$$
\begin{align*}
& \binom{s_{2}}{n_{2}} \sum_{j_{1}=0}^{n_{1}}\left(\frac{\mu_{1}}{\lambda_{2}}\right)^{j_{1}}\binom{s_{1}-j_{1}}{n_{1}-j_{1}}+\binom{s_{1}}{n_{1}} \sum_{j_{2}=0}^{n_{2}}\left(\frac{\mu_{2}}{\lambda_{1}}\right)^{j_{2}}\binom{s_{2}-j_{2}}{n_{2}-j_{2}}-\left(\begin{array}{l}
\underset{\sim}{\tilde{n}}
\end{array}\right)  \tag{2.12}\\
& =\sum_{\underset{\sim}{k} \geq 0}\binom{r_{1}(\underset{\sim}{k})}{k_{1}}\binom{r_{2}(\underset{\sim}{k})}{k_{2}}\binom{-r_{1}(\underset{\sim}{k})+s_{1}}{n_{1}-k_{1}}\binom{-r_{2}(\underset{\sim}{k})+s_{2}}{n_{2}-k_{2}} .
\end{align*}
$$

$$
\begin{align*}
& \frac{s_{2}^{n_{2}}}{n_{2}!} \sum_{j_{1}=0}^{n_{1}}\left(\frac{\mu_{1}}{\lambda_{2}}\right)^{\dot{j}_{1}} \frac{s_{1}^{n_{1}-j_{1}}}{\left(n_{1}-j_{1}\right)!}+\frac{s_{1}^{n_{1}}}{n_{1}!} \sum_{j_{2}=0}^{n_{2}}\left(\frac{\mu_{2}}{\lambda_{1}}\right)^{j_{2}} \frac{s_{2}^{n_{2}-j_{2}}}{\left(n_{2}-j_{2}\right)!}-\frac{s^{n}}{\underset{\sim}{n}!}  \tag{2.11}\\
& =\sum_{\underset{\sim}{k} \geq} \frac{\left.r_{\sim}^{r}(\underset{\sim}{k})^{k_{1}}{\underset{r}{2}}^{(\underset{\sim}{k}}\right)^{k_{2}}}{\underset{\sim}{k}!} \frac{\left(-r_{1}(\underset{\sim}{k})+s_{1}\right)^{n_{1}-k_{1}} \cdot\left(-r_{2}(\underset{\sim}{k})+s_{2}\right)^{n_{2}-k_{2}}}{(\underset{\sim}{n}-\underset{\sim}{k})!} .
\end{align*}
$$

$$
f(\underset{\sim}{z})=\left(z_{1} / \phi_{1}(\underset{\sim}{z}), z_{2} / \phi_{2}(\underset{\sim}{z})\right) .
$$

The (two-variable) Lagrange-Good formula solves the problem of expanding a formal Laurent series $g(z)$ of the form

$$
\begin{equation*}
g(\underset{\sim}{z})=\sum_{\underset{\sim}{j} \geq \underset{\sim}{m}} g_{\sim}^{j} \underset{\sim}{z} \underset{\sim}{j}, \tag{3.1}
\end{equation*}
$$

for some $\underset{\sim}{m} \in \mathbb{Z}^{2}$, in terms of powers of $f(z)$; namely, if

$$
g(\underset{\sim}{z})=\sum_{\underset{\sim}{x} \in \mathbb{Z}^{2}} c_{\underset{\sim}{k}} f_{\sim}^{k}(\underset{\sim}{z}),
$$

then

$$
c_{\underset{\sim}{k}}=\left\langle{\underset{\sim}{z}}^{0}\right\rangle g(\underset{\sim}{z}) \Delta(\underset{\sim}{z}) f^{-k}(\underset{\sim}{z}),
$$

where $\langle\underset{\sim}{z}\rangle \alpha(\underset{\sim}{z})$ denotes the coefficient of $\underset{\sim}{\underset{\sim}{z}} 0$ in $\alpha(\underset{\sim}{z})$, and

$$
\Delta(\underset{\sim}{z})=\frac{\partial f}{\partial z}(\underset{\sim}{z}) \phi_{1}(\underset{\sim}{z}) \phi_{2}(\underset{\sim}{z}) \text { with } \frac{\partial f}{\partial z}(\underset{\sim}{z}) \text { the Jacobian of } f(\underset{\sim}{z}) .
$$

For formal Laurent series of the form (3.1), we shall use the abbreviation fLs. The general two-dimensional Lagrange inversion problem can be formulated as follows: Let $\left.F=\left(f_{\underset{\sim}{k}}^{(\underset{\sim}{z}}\right)\right)_{\underset{\sim}{k} \in \mathbb{Z}^{2}}$ be a "diagonal sequence," i.e., $\left.f_{\underset{\sim}{k}}^{\underset{\sim}{z}} \underset{\sim}{z}\right)$ is of the form

$$
f_{\underset{\sim}{k}}(\underset{\sim}{z})=\sum_{\underset{\sim}{n} \geq \underset{\sim}{k}} f_{\sim \sim}^{n k}{\underset{\sim}{z}}_{\underset{\sim}{k}} .
$$

Then, for a given sequence $F$, one tries to find some sequence $\tilde{F}=\left(\tilde{f}_{\underset{\sim}{k}}(\underset{\sim}{z})\right)_{\underset{\sim}{k} \in \mathbb{Z}^{2}}$ such that expanding an arbitrary $f L s g(\underset{\sim}{z})$ in terms of $F$,
(3.2a) $g(\underset{\sim}{z})=\sum_{\underset{\sim}{k} \in \mathbf{Z}^{2}}{\underset{\sim}{\underset{\sim}{*}}}^{\sim} f_{\underset{\sim}{k}}(\underset{\sim}{z})$,
the coefficients ${\underset{\sim}{k}}^{\sim}$ are given by

Obviously, the sequence $\tilde{\tilde{F}}$ is uniquely determined by
(3.3) $\left\langle{\underset{\sim}{\sim}}^{0}\right\rangle{\underset{\sim}{k}}_{\underset{\sim}{x}}(\underset{\sim}{z}) \cdot \tilde{f}_{\underset{\sim}{n}}(\underset{\sim}{z})=\delta_{\underset{\sim}{n k}}$,
where $\delta_{n k}^{n}$ is the Kronecker delta. In this paper, we shall solve this Lagrange inversion problem for
(3.4) $f_{\underset{\sim}{k}}(\underset{\sim}{z})={\underset{\sim}{z}}_{\sim}^{k} f_{1}\left(z_{1}\right)^{R_{1}(\underset{\sim}{k})} f_{2}\left(z_{2}\right)^{R_{2}(\underset{\sim}{k})}$,
where $f_{1}(t), f_{2}(t)$ are $f p s$ in the single variable $t$ with $f_{i}(0) \neq 0, i=1,2$. Evidently, $\left.F=\left(f_{\underset{k}{k}}^{(\underset{\sim}{z}}\right)\right)_{\underset{\sim}{c} \in \mathbf{Z}^{2}}$ is a diagonal sequence.
Theorem 1: Let $\left.F=\left({\underset{\sim}{\underset{\sim}{k}}}^{\underset{\sim}{z}} \underset{\sim}{z}\right)\right)_{\underset{\sim}{k} \in \mathbf{Z}^{2}}$ be as defined in (3.4). The sequence

$$
\tilde{F}=\left(\tilde{f}_{\underset{\sim}{k}}^{\underset{\sim}{z}}(\underset{\sim}{z})\right)_{\underset{\sim}{k} \in \mathbf{Z}^{2}}^{\sim},
$$

uniquely determined by (3.3), is given by

$$
\text { (3.5) } \left.\quad \underset{f_{\underset{\sim}{k}}}{\underset{\sim}{z}}\right)=\frac{\mu_{1} \mu_{2}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)} W f_{\underset{\sim}{k}}(\underset{\sim}{z})^{-1} \text {, }
$$

where $W$ is the operator
(3.6) $\quad W=\operatorname{det}\left|\begin{array}{lr}-z_{1} D_{1}+\frac{\lambda_{1}}{\mu_{1}} & \left(\alpha_{2}-\frac{\beta_{2}}{\mu_{2}} \lambda_{2}\right) F_{1}\left(z_{1}\right) \\ \left(\alpha_{1}-\frac{\beta_{1}}{\mu_{1}} \lambda_{1}\right) F_{2}\left(z_{2}\right) & -z_{2} D_{2}+\frac{\lambda_{2}}{\mu_{2}}\end{array}\right|$
with

$$
F_{i}\left(z_{i}\right)=z_{i} \frac{\partial f_{i}}{\partial z_{i}}\left(z_{i}\right) / f_{i}\left(z_{i}\right), i=1,2 .
$$

$D_{i}$ stands for the differential operator with respect to $z_{i}$. Equivalently,

$$
\begin{align*}
\underset{\sim}{\tilde{f}_{\underset{k}{ }}(z)}=[(1 & \left.+\mu_{1} \frac{\alpha_{2}+\beta_{2} k_{2}}{\lambda_{2}+\mu_{2} k_{2}} F_{1}\left(z_{1}\right)\right)\left(1+\mu_{2} \frac{\alpha_{1}+\beta_{1} k_{1}}{\lambda_{1}+\mu_{1} k_{1}} F_{2}\left(z_{2}\right)\right)  \tag{3.7}\\
& \left.-\frac{\left(\alpha_{1} \mu_{1}-\beta_{1} \lambda_{1}\right)\left(\alpha_{2} \mu_{2}-\beta_{2} \lambda_{2}\right)}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)} F_{1}\left(z_{1}\right) F_{2}\left(z_{2}\right)\right] \underset{\sim}{{\underset{\sim}{k}}(z)^{-1} .}
\end{align*}
$$

Proof: The proof is based on the method for treating Lagrange inversion problems introduced by the author [11]. For $i=1,2$, apply ( $z_{i} D_{i}+\left(\lambda_{i} / \mu_{i}\right)$ ) on both sides of (3.4) to get

$$
\begin{align*}
& \left(z_{i} D_{i}+\frac{\lambda_{i}}{\mu_{i}}\right) f_{\underset{\sim}{k}}(\underset{\sim}{z})  \tag{3.8}\\
& =\left[\frac{\lambda_{i}+\mu_{i} k_{i}}{\mu_{i}}+\left(\lambda_{i}+\mu_{i} k_{i}\right)\left(\frac{\beta_{3-i}}{\mu_{3-i}}+\frac{\alpha_{3-i}-\frac{\beta_{3-i}}{\mu_{3-i}} \lambda_{3-i}}{\lambda_{3-i}+\mu_{3-i} k_{3-i}}\right) F_{i}\left(z_{i}\right)\right] \underset{\sim}{f_{k}}(\underset{\sim}{z}) \\
& i=1,2 .
\end{align*}
$$

Writing $c_{i}(\underset{\sim}{k})=\left(\lambda_{i}+\mu_{i} k_{i}\right)^{-1}, i=1,2$, simple manipulations show that the system (3.8) of two equations is equivalent to the system

$$
\begin{equation*}
U_{i} f_{\underset{\sim}{k}}(\underset{\sim}{z})=c_{i}(\underset{\sim}{k}) V f_{\underset{\sim}{k}}(\underset{\sim}{z}), i=1,2, \tag{3.9}
\end{equation*}
$$

where

$$
\begin{aligned}
U_{i}= & \left(z_{3-i} D_{3-i}+\frac{\lambda_{3-i}}{\mu_{3-i}}\right)\left(\frac{1}{\mu_{i}}+\frac{\beta_{3-i}}{\mu_{3-i}} F_{i}\left(z_{i}\right)\right) \\
& +\left(\alpha_{3-i}-\frac{\beta_{3-i}}{\mu_{3-i}} \lambda_{3-i}\right) F_{i}\left(z_{i}\right)\left(\frac{1}{\mu_{3-i}}+\frac{\beta_{i}}{\mu_{i}} F_{3-i}\left(z_{3-i}\right)\right)
\end{aligned}
$$

and

$$
V=\operatorname{det}\left|\begin{array}{lr}
z_{1} D_{1}+\frac{\lambda_{1}}{\mu_{1}} & \left(\alpha_{2}-\frac{\beta_{2}}{\mu_{2}} \lambda_{2}\right) F_{1}\left(z_{1}\right)  \tag{3.10}\\
\left(\alpha_{1}-\frac{\beta_{1}}{\mu_{1}} \lambda_{1}\right) F_{2}\left(z_{2}\right) & z_{2} D_{2}+\frac{\lambda_{2}}{\mu_{2}}
\end{array}\right|
$$

Now Theorem 1 of [11] with $A=\mathbb{C}, M_{1}=M_{2}=$ set of fLs, $U_{i}, V, c_{i}(\underset{\sim}{k})$ as above, may be applied. The bilinear form we need is defined by
(3.11) $(a(\underset{\sim}{z}), b(\underset{\sim}{z}))=\langle\underset{\sim}{z} \xlongequal{0}\rangle a(\underset{\sim}{z}) b(\underset{\sim}{z})$,
for fLs $a(\underset{\sim}{z})$ and $b(\underset{\sim}{z})$. Thus, by (4.4) of [11], the dual system
(3.12) $U_{i}^{*} h_{\underset{\sim}{k}}(\underset{\sim}{z})=c_{i}(\underset{\sim}{k}) W h_{\sim}^{k}(\underset{\sim}{z}), i=1,2$,
[note that $\tilde{V}^{*}=W$ as defined in (3.6), since $\left(z_{i} D_{i}\right)^{*}=-z_{i} D_{i}$ ] has to be solved first. It is a simple matter of fact that (3.12), the dual of (3.9), is equivalent to the dual of (3.8), which reads

$$
\begin{aligned}
& \left(-z_{i} D_{i}+\frac{\lambda_{i}}{\mu_{i}}\right) h_{\underset{\sim}{k}}(\underset{\sim}{z}) \\
& =\left[\frac{\lambda_{i}+\mu_{i} k_{i}}{\mu_{i}}+\left(\lambda_{i}+\mu_{i} k_{i}\right)\left(\frac{\beta_{3-i}}{\mu_{3-i}}+\frac{\alpha_{3-i}-\frac{\beta_{3-i}}{\mu_{3-i}} \lambda_{3-i}}{\lambda_{3-i}+\mu_{3-i} k_{3-i}}\right) F_{i}\left(z_{i}\right)\right] \underset{\sim}{h_{k}}(\underset{\sim}{z}) \\
& i=1,2 .
\end{aligned}
$$

A solution of this system of equations is seen to be $h_{\underset{\sim}{k}}(\underset{\sim}{z})=\underset{\sim}{f} \underset{\sim}{f}(\underset{\sim}{z})^{-1}$, hence, by (4.6) of [11], respecting $V^{*}=W$,

$$
\tilde{f}_{\underset{\sim}{k}}(\underset{\sim}{z})=\frac{\mu_{1} \mu_{2}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)} W h_{\underset{\sim}{x}}^{\underset{\sim}{z}}(\underset{\sim}{z}),
$$

which establishes (3.5). A little bit of calculation from (3.5) leads to (3.7) .

Corollary 2 (Lagrange formula) : Let $F=\left(f_{\underset{\sim}{k}}(\underset{\sim}{z})\right)_{\underset{\sim}{k} \in \mathbf{Z}^{2}}$ be as defined in (3.4). The coefficients in the expansion

$$
\begin{align*}
& g(\underset{\sim}{z})=\sum_{\underset{\sim}{k} \in \mathbb{Z}^{2}}{\underset{\sim}{x}}_{\underset{\sim}{k}} f_{\sim}^{k}(\underset{\sim}{z})  \tag{3.13}\\
& \text { en })
\end{align*}
$$

are given by
(3.14) $\quad c_{\underset{\sim}{k}}=\langle\underset{\sim}{z} \underset{\sim}{0}\rangle g(\underset{\sim}{z}) \underset{\sim}{\underset{\sim}{k}} \underset{\sim}{z}(\underset{\sim}{z})$,
with $\underset{f_{k}}{\underset{\sim}{x}}(\underset{\sim}{z})$ of $(3.7)$, or
(3.15) $\quad \underset{\sim}{k}=\frac{\mu_{1} \mu_{2}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)}\left\langle\underset{\sim}{\underset{\sim}{\sim}} \underset{\sim}{0}{\underset{\sim}{k}}_{\underset{\sim}{k}}(\underset{\sim}{z})^{-1} V g(\underset{\sim}{z})\right.$.
\left. Proof: Equation (3.14) is merely (3.2) for ${\underset{\sim}{\underset{\sim}{k}}}_{\underset{\sim}{~}}^{\underset{\sim}{z}}\right)$ of $(3.4)$, (3.15) is based on (3.14), (3.5), and $W^{*}=V$ 。

As a first application, we shall prove (1.7) of Cohen \& Hudson [5]. Take $f_{1}(t)=f_{2}(t)=\exp t$, which implies $F_{1}(t)=F_{2}(t)=t$. Let

$$
g(\underset{\sim}{z})=\frac{1}{\lambda_{1} \lambda_{2}} \sum_{j=0}^{\infty} \frac{\left[\left(\alpha_{1}-\frac{\beta_{1}}{\mu_{1}} \lambda_{1}\right)\left(\alpha_{2}-\frac{\beta_{2}}{\mu_{2}} \lambda_{2}\right) z_{1} z_{2}\right]^{j}}{\left(\lambda_{1} / \mu_{1}+1\right)_{j}\left(\lambda_{2} / \mu_{2}+1\right)_{j}}
$$

where $(\alpha)_{j}=\alpha(\alpha+1) \ldots(\alpha+j-1)$. For this choice of $f_{i}(t)$ [ $V$ depends on $\left.F_{i}(t)!\right], g(\underset{\sim}{z})$ satisfies $V g(\underset{\sim}{z})=1 / \mu_{1} \mu_{2}$. Utilizing the Lagrange formula (3.15), from this fact we obtain

$$
\begin{align*}
& \frac{1}{\lambda_{1} \lambda_{2}} \sum_{j=0}^{\infty} \frac{\left[\left(\alpha_{1}-\frac{\beta_{1}}{\mu_{1}} \lambda_{1}\right)\left(\alpha_{2}-\frac{\beta_{2}}{\mu_{2}} \lambda_{2}\right) z_{1} z_{2}\right]^{j}}{\left(\lambda_{1} / \mu_{1}+1\right)_{j}\left(\lambda_{2} / \mu_{2}+1\right)_{j}}  \tag{3.16}\\
& =\sum_{\underset{\sim}{k} \geq \underset{\sim}{0}}\left(\lambda_{1}+\mu_{1} k_{1}\right)^{-1}\left(\lambda_{2}+\mu_{2} k_{2}\right)^{-1} \frac{R_{1}(\underset{\sim}{k})^{k_{1}} R_{2}(\underset{\sim}{k})^{k_{2}}}{\underset{\sim}{k}!}(-\underset{\sim}{z})_{\sim}^{k} \exp \left(R_{1}(\underset{\sim}{k}) z_{1}+R_{2}(\underset{\sim}{k}) z_{2}\right) .
\end{align*}
$$

Equation (3.16) is another two-variable extension of (1.1) (set $z_{2}=0$ ).

## 4. Coefficient Formulas for Some Special Expansions

The following technical lemma turns out to be useful for further computations.
Lemma 3: Let $h(t)$ be an fLs in $t$. With the assumptions of Theorem 1 ,

$$
\begin{align*}
& \langle\underset{\sim}{z} \stackrel{0}{\sim}\rangle h\left(z_{3-i}\right) F_{i}\left(z_{i}\right) f_{\underset{\sim}{k}}(\underset{\sim}{z})^{-1}  \tag{4.1}\\
& =-\frac{k_{i}\left(\lambda_{3-i}+\mu_{3-i} k_{3-i}\right)}{\left(\lambda_{i}+\mu_{i} k_{i}\right)\left(\alpha_{3-i}+\beta_{3-i} k_{3-i}\right)}\langle\underset{\sim}{z} \stackrel{0}{\sim}\rangle h(z 3-i) f_{\underset{\sim}{k}}(\underset{\sim}{z})^{-1} \text {, for } i=1,2 \text {. }
\end{align*}
$$

Proof: Without loss of generality we prove (4.1) for $i=1$. We start with the identity

$$
\begin{aligned}
&\langle\underset{\sim}{z}\underset{\sim}{0}\rangle \\
&\left(z_{2}\right) \\
&=\left.-\langle\underset{\sim}{z} \underset{\sim}{0}\rangle h\left(z_{1}\right) \frac{\alpha_{2}+\beta_{2} k_{2}}{\lambda_{2}+\mu_{2} k_{2}} F_{1}\left(z_{1}\right)\right){\underset{\sim}{k}}_{\underset{\sim}{k}}(\underset{\sim}{z})^{-1} \\
& \cdot \frac{1}{k_{1}} f_{2}\left(z_{2}\right)^{-R_{2}(\underset{\sim}{k})} f_{1}\left(z_{1}\right)^{-\lambda_{1}}\left(\alpha_{2}+z_{2} k_{2}\right) /\left(\lambda_{2}+\mu_{2} k_{2}\right)
\end{aligned} .
$$

Because of $\left(z_{1} D_{1}\right)^{*}=-z_{1} D_{1}$ [with respect to the bilinear form of (3.11)] the right-hand side of this equation is equal to

$$
\begin{gathered}
\frac{1}{k_{1}}\langle\underset{\sim}{\sim} \stackrel{0}{\sim}\rangle z_{1} D_{1}\left[h\left(z_{2}\right) z_{2}^{-k_{2}} f_{2}\left(z_{2}\right)^{-R_{2}(\underset{\sim}{k})} f_{1}\left(z_{1}\right)^{-\lambda_{1}\left(\alpha_{2}+\beta_{2} k_{2}\right) /\left(\lambda_{2}+\mu_{2} k_{2}\right)}\right] \cdot \\
\cdot z_{1}^{-k_{1}} f_{1}\left(z_{1}\right)^{-\mu_{1} k_{1}\left(\alpha_{2}+\beta_{2} k_{2}\right) /\left(\lambda_{2}+\mu_{2} k_{2}\right)}
\end{gathered}
$$

Together with a bit of manipulation, we finally arrive at (4.1). Corollary 4: Let $f_{k}(\underset{\sim}{z})$ be given by (3.4). Then
(4.2) $1=\sum_{\underset{\sim}{k} \geq \underset{\sim}{0}} \frac{\lambda_{1} \lambda_{2}-k_{1} k_{2} \frac{\left(\alpha_{1} \mu_{1}-\beta_{1} \lambda_{1}\right)\left(\alpha_{2} \mu_{2}-\beta_{2} \lambda_{2}\right)}{\left(\alpha_{1}+\beta_{1} k_{1}\right)\left(\alpha_{2}+\beta_{2} k_{2}\right)}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)}{\underset{\sim}{\underset{\sim}{k}}}^{f_{\underset{\sim}{*}}^{\underset{\sim}{k}}(\underset{\sim}{z})}$,
where $\left.d_{\underset{\sim}{k}}=\langle\underset{\sim}{z} \underset{\sim}{0}\rangle{\underset{\sim}{k}}^{\underset{\sim}{z}} \underset{\sim}{z}\right)^{-1}$.
Proof: By (3.14) we have to compute

$$
\langle\underset{\sim}{z}\rangle 1 \cdot \underset{\sim}{\tilde{f}} \underset{\sim}{k}(\underset{\sim}{z})=\langle\underset{\sim}{z} \underset{\sim}{0}\rangle \tilde{f}_{k}(\underset{\sim}{z}) .
$$

Using the form (3.7) of $\tilde{f}_{k}(\underset{\sim}{z})$, repeated application of (4.1) gives

$$
\begin{aligned}
& \langle\underset{\sim}{\underset{\sim}{\sim}}\rangle \underset{\sim}{\underset{f}{k}} \underset{\sim}{z} \underset{\sim}{z})=\left\langle\underset { \sim } { \underset { \sim } { \sim } } { } _ { \sim } ^ { \sim } \left[\left(1-\mu_{1} \frac{k_{1}}{\lambda_{1}+\mu_{1} k_{1}}\right)\left(1-\mu_{2} \frac{k_{2}}{\lambda_{2}+\mu_{2} k_{2}}\right)\right.\right. \\
& \left.-\frac{\left(\alpha_{1} \mu_{1}-\beta_{1} \lambda_{1}\right)\left(\alpha_{2} \mu_{2}-\beta_{2} \lambda_{2}\right)}{\left(\alpha_{1}+\beta_{1} k_{1}\right)\left(\alpha_{2}+\beta_{2} k_{2}\right)} \cdot \frac{k_{1} k_{2}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)}\right] f_{\mathcal{K}}(\underset{\sim}{z})^{-1} \\
& =\frac{\lambda_{1} \lambda_{2}-k_{1} k_{2} \frac{\left(\alpha_{1} \mu_{1}-\beta_{1} \lambda_{1}\right)\left(\alpha_{2} \mu_{2}-\beta_{2} \lambda_{2}\right)}{\left(\alpha_{1}+\beta_{1} k_{1}\right)\left(\alpha_{2}+\beta_{2} k_{2}\right)}}{\left(\lambda_{1}+\mu_{1} k_{1}\right)\left(\lambda_{2}+\mu_{2} k_{2}\right)}\langle\underset{\sim}{z}\rangle \underset{\sim}{\underset{\sim}{k}} \underset{\sim}{z}(\underset{\sim}{r})^{-1},
\end{aligned}
$$

which furnishes (4.2). $\square$
For $f_{i}\left(z_{i}\right)=\exp \left(-z_{i}\right)$ and $f_{i}\left(z_{i}\right)=\left(1+z_{i}\right)^{-1}$, respectively, the expansions (2.8) and (2.9) are obtained as special cases of (4.2). The mixed analogue (2.10) is (4.2) with $f_{1}\left(z_{1}\right)=\exp \left(-z_{1}\right)$ and $f_{2}\left(z_{2}\right)=\left(1+z_{2}\right)^{-1}$.

Quite analogously, we prove
Corollary 5: If $f_{k}(\underset{\sim}{z})=\underset{\sim}{z} \underset{\sim}{k} f_{1}\left(z_{1}\right)^{r_{1}(\underset{\sim}{k})} f_{2}\left(z_{2}\right)^{R_{2}(\underset{\sim}{k})}$, then

$$
\begin{equation*}
\frac{1}{\lambda_{1}+\mu_{2} \alpha_{1} F_{2}\left(z_{2}\right)}=\sum_{\underset{\sim}{k} \geq \underset{\sim}{0}}\left(\lambda_{1}+\mu_{1} k_{1}\right)^{-1} d_{\underset{\sim}{k}}^{f_{\sim}^{k}} \underset{\sim}{\underset{\sim}{z}}(\underset{\sim}{z}), \tag{4.3}
\end{equation*}
$$

where $\left.d_{\underset{\sim}{k}}=\langle\underset{\sim}{z} \underset{\sim}{0}\rangle \underset{\sim}{\underset{\sim}{k}} \underset{\sim}{z}\right)^{-1}$.
Sketch of Proof: Again using (3.14), (3.7), and (4.1) we proceed along the lines of the proof of the preceding corollary.

The expansions (2.4)-(2.7) are special cases of (4.3). Finally, we have Corollary 6: If $\left.f_{\underset{\sim}{k}}^{\underset{\sim}{z}} \underset{\sim}{z}\right)=\underset{\sim}{\underset{\sim}{\sim}} \underset{\sim}{k} f_{1}\left(z_{1}\right)^{r_{1}(\underset{\sim}{k})} f_{2}\left(z_{2}\right)^{r_{2}(\underset{\sim}{k})}$, then
(4.4) $\frac{1-\frac{\mu_{1} \mu_{2}}{\lambda_{1} \lambda_{2}} F_{1}\left(z_{1}\right) F_{2}\left(z_{2}\right)}{\left(1+\frac{\mu_{1}}{\lambda_{2}} F_{1}\left(z_{1}\right)\right)\left(1+\frac{\mu_{2}}{\lambda_{1}} F_{2}\left(z_{2}\right)\right)}=\sum_{\underset{\sim}{k} \geq \underset{\sim}{0}}^{\left(d_{\sim}^{k}\right.} f_{\sim} \underset{\sim}{k}(\underset{\sim}{z})$,
where $d_{\underset{\sim}{k}}=\langle\underset{\sim}{z} \underset{\sim}{0}\rangle{\underset{\sim}{f}}_{\underset{\sim}{k}}(\underset{\sim}{z})^{-1}$.

Proof: Observe that the left-hand side of (4.4) is equal to

$$
1 /\left(1+\frac{\mu_{1}}{\lambda_{2}} F_{1}\left(z_{1}\right)\right)+1 /\left(1+\frac{\mu_{2}}{\lambda_{1}} F_{2}\left(z_{2}\right)\right)-1
$$

This in hand, the method used to prove Corollary 4 can be used again to settle (4.4) .

Equations (2.1)-(2.3) are special cases of (4.4).

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# LAFMBERT SERIES AND THE SUMMATION OF RECIPROCALS IN CERTAIN FIBONACCI-LUCAS-TYPE SEQUENCES 

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1. Introduction

Consider the sequence of real numbers defined by the recurrence relation (1.1) $\quad W_{n}=p W_{n-1}+W_{n-2}$,
where $p$ is a strictly positive real number. Special cases of $\left(W_{n}\right)$ which interest us here are:
(1.2) $\quad U_{n}=\frac{\alpha^{n}-\beta^{n}}{\alpha-\beta}$ (Fibonacci-type sequence),
and
(1.3) $\quad V_{n}=\alpha^{n}+\beta^{n}$ (Lucas-type sequence),
where $\quad \alpha=\frac{p+\sqrt{p^{2}+4}}{2}$,
(1.4)

$$
\beta=\frac{p-\sqrt{p^{2}+4}}{2} .
$$

It is clear that
(1.5) $\alpha \beta=-1, \alpha>1,-1<\beta<0$.

On the other hand, the Lambert series is defined by
(1.6) $L(x)=\sum_{n=1}^{\infty} \frac{x^{n}}{1-x^{n}}, \quad|x|<1$.

It has been known for a long time (see Horadam [1] for complete references) that

$$
\begin{aligned}
& \sum_{n=1}^{\infty} \frac{1}{U_{2 n}}=(\alpha-\beta)\left[L\left(\beta^{2}\right)-L\left(\beta^{4}\right)\right] \\
& \sum_{n=1}^{\infty} \frac{1}{V_{2 n}-1}=-L(\beta)+2 L\left(\beta^{2}\right)-L\left(\beta^{4}\right) .
\end{aligned}
$$

The purpose of this paper is to establish the following result.
Theorem 1:
(1.7) $\sum_{n=1}^{\infty} \frac{1}{U_{n} U_{n+1}}=2(\alpha-\beta)\left[L\left(\beta^{2}\right)-2 L\left(\beta^{4}\right)+2 L\left(\beta^{8}\right)\right]+\beta$;
(1.8) $\quad \sum_{n=1}^{\infty} \frac{1}{V_{n} V_{n}+1}=\frac{2}{\alpha-\beta}\left[L\left(\beta^{2}\right)-2 L\left(\beta^{8}\right)\right]+\frac{\beta}{(\alpha-\beta) p}$.
2. Preliminary Lemma

Lemma 1:
(2.1) $\sum_{n=0}^{\infty} \frac{x^{2 n+1}}{1-x^{2 n+1}}=L(x)-L\left(x^{2}\right)$;

Lambert series and the summation of reciprocals in certain fibonacci-Lucas-TYPe sequences
(2.2) $\sum_{n=1}^{\infty} \frac{x^{n}}{1+x^{n}}=L(x)-2 L\left(x^{2}\right)$;
(2.3) $\sum_{n=0}^{\infty} \frac{x^{2 n+1}}{1+x^{2 n+1}}=L(x)-3 L\left(x^{2}\right)+2 L\left(x^{4}\right)$.

$$
\begin{aligned}
& \text { (2.1) is obviously true, whereas (2.2) follows from the identity } \\
& \frac{x^{n}}{1+x^{n}}=\frac{x^{n}}{1-x^{n}}-\frac{2 x^{2 n}}{1-x^{2 n}}
\end{aligned}
$$

and (2.3) follows from

$$
\sum_{n=0}^{\infty} \frac{x^{2 n+1}}{1+x^{2 n+1}}=\sum_{n=1}^{\infty} \frac{x^{n}}{1+x^{n}}-\sum_{n=1}^{\infty} \frac{x^{2 n}}{1+x^{2 n}}
$$

## 3. Proof of Theorem 1

Lemma 2:

$$
\begin{align*}
& 2 \sum_{n=1}^{\infty} \frac{1}{\alpha^{n} U_{n}}=\frac{1}{\alpha}+\sum_{n=1}^{\infty} \frac{1}{U_{n} U_{n}+1}  \tag{3.1}\\
& 2 \sum_{n=1}^{\infty} \frac{1}{\alpha^{n} V_{n}}=\frac{1}{\alpha p}+\sum_{n=1}^{\infty} \frac{\alpha-\beta}{V_{n} V_{n}+1} \tag{3.2}
\end{align*}
$$

Proof: First, we have

$$
\begin{aligned}
\alpha U_{n+1}+U_{n} & =\frac{1}{\alpha-\beta}\left[\alpha\left(\alpha^{n+1}-(-1)^{n+1} \frac{1}{\alpha^{n+1}}\right)+\alpha^{n}-(-1)^{n} \frac{1}{\alpha^{n}}\right] \\
& =\frac{1}{\alpha-\beta}\left(\alpha^{n+2}+\alpha^{n}\right)=\frac{\alpha^{n+1}}{\alpha-\beta}\left(\alpha+\frac{1}{\alpha}\right)=\alpha^{n+1} .
\end{aligned}
$$

Thus,

$$
\frac{1}{\alpha^{n} U_{n}}+\frac{1}{\alpha^{n+1} U_{n+1}}=\frac{1}{U_{n} U_{n+1}} .
$$

By adding this term by term, we find (3.1) since $U_{1}=1$. The proof of (3.2) follows the same pattern if we observe that

$$
\alpha V_{n+1}+V_{n}=(\alpha-\beta) \alpha^{n+1}
$$

Thus,

$$
\frac{1}{\alpha^{n} V_{n}}+\frac{1}{\alpha^{n+1} V_{n+1}}=\frac{\alpha-\beta}{V_{n} V_{n+1}} .
$$

Now, adding this term by term, we find (3.2) since $V_{1}=p$.
Lemma 3:

$$
\begin{equation*}
\sum_{n=1}^{\infty} \frac{1}{\alpha^{n} U_{n}}=(\alpha-\beta)\left[L\left(\beta^{2}\right)-2 L\left(\beta^{4}\right)+2 L\left(\beta^{8}\right)\right] ; \tag{3.3}
\end{equation*}
$$

(3.4) $\sum_{n=1}^{\infty} \frac{1}{\alpha^{n} V_{n}}=L\left(\beta^{2}\right)-2 L\left(\beta^{8}\right)$.

Proof:

$$
\begin{aligned}
\frac{1}{\alpha-\beta} \sum_{n=1}^{\infty} \frac{1}{\alpha^{n} U_{n}} & =\sum_{n=1}^{\infty} \frac{1}{\alpha^{2 n}-(-1)^{n}}=\sum_{n=1}^{\infty} \frac{\beta^{2 n}}{1-(-1)^{n} \beta^{2 n}} \\
& =\sum_{n=1}^{\infty} \frac{\beta^{4 n}}{1-\beta^{4 n}}+\sum_{n=0}^{\infty} \frac{\beta^{4 n+2}}{1+\beta^{4 n+2}}
\end{aligned}
$$

Using (1.6) with $x=\beta^{4}$ and (2.3) with $x=\beta^{2}$, we find (3.3). On the other hand, we have:

LAMBERT SERIES AND THE SUMMATION OF RECIPROCALS IN CERTAIN FIBONACCI-LUCAS-TYPE SEQUENCES

$$
\sum_{n=1}^{\infty} \frac{1}{\alpha^{n} V_{n}}=\sum_{n=1}^{\infty} \frac{1}{\alpha^{2 n}+(-1)^{n}}=\sum_{n=1}^{\infty} \frac{\beta^{2 n}}{1+(-1)^{n} \beta^{2 n}}=\sum_{n=1}^{\infty} \frac{\beta^{4 n}}{1+\beta^{4 n}}+\sum_{n=0}^{\infty} \frac{\beta^{4 n+2}}{1-\beta^{4 n+2}}
$$

Using (2.2) with $x=\beta^{4}$ and (2.1) with $x=\beta^{2}$, we find (3.4). This concludes the proof of Lemma 3. Now the proof of the theorem follows immediately from Lemmas 2 and 3.

## 4. Special Cases

### 4.1 Fibonacci-Lucas Sequences

Let $p=1$ in (1.1) to obtain

$$
W_{n}=W_{n-1}+W_{n-2}, \quad \alpha=\frac{1+\sqrt{5}}{2}, \quad \beta=\frac{1-\sqrt{5}}{2} .
$$

$U_{n}=F_{n}$ is the Fibonacci sequence and $V_{n}=L_{n}$ is the Lucas sequence. Equations (1.7) and (1.8) take the following form:

$$
\begin{aligned}
& \sum_{n=1}^{\infty} \frac{1}{F_{n} F_{n+1}}=2 \sqrt{5}\left[L\left(\frac{3-\sqrt{5}}{2}\right)-2 L\left(\frac{7-3 \sqrt{5}}{2}\right)+2 L\left(\frac{47-21 \sqrt{5}}{2}\right)\right]+\frac{1-\sqrt{5}}{2} ; \\
& \sum_{n=1}^{\infty} \frac{1}{L_{n} L_{n}+1}=\frac{2}{\sqrt{5}}\left[L\left(\frac{3-\sqrt{5}}{2}\right)-2 L\left(\frac{47-21 \sqrt{5}}{2}\right)\right]+\frac{1-\sqrt{5}}{2 \sqrt{5}}
\end{aligned}
$$

### 4.2 Pell and Pell-Lucas Sequences

Let $p=2$ in (1.1) to obtain
$W_{n}=2 W_{n-1}+W_{n-2}, \quad \alpha=1+\sqrt{2}, \quad \beta=1-\sqrt{2}$.
$U_{n}=P_{n}$ is the Pell sequence, $V_{n}=Q_{n}$ is the Pell-Lucas sequence. Equations (1.7) and (1.8) take the form:

$$
\begin{aligned}
& \sum_{n=1}^{\infty} \frac{1}{P_{n} P_{n+1}}=4 \sqrt{2}[L(3-2 \sqrt{2})-2 L(17-12 \sqrt{2})+2 L(577-408 \sqrt{2})]+1-\sqrt{2} \\
& \sum_{n=1}^{\infty} \frac{1}{Q_{n} Q_{n}+1}=\frac{1}{\sqrt{2}}[L(3-2 \sqrt{2})-2 L(577-408 \sqrt{2})]+\frac{1-\sqrt{2}}{4 \sqrt{2}}
\end{aligned}
$$

## 5. Generalization

The following theorem generalizes the above result. It is given without proof, since the methods required exactly parallel those of Section 3 . We assume that $K$ is an odd integer.
Theorem 2:

$$
\begin{aligned}
& \sum_{n=1}^{\infty} \frac{1}{U_{k n} U_{k(n+1)}}=\frac{2(\alpha-\beta)}{U_{k}}\left[L\left(\beta^{2 k}\right)-2 L\left(\beta^{4 k}\right)+2 L\left(\beta^{8 k}\right)\right]+\frac{\beta^{k}}{U_{k}^{2}} \\
& \sum_{n=1}^{\infty} \frac{1}{V_{k n} V_{k(n+1)}}=\frac{2}{(\alpha-\beta) U_{k}}\left[L\left(\beta^{2 k}\right)-2 L\left(\beta^{8 k}\right)\right]+\frac{\beta^{k}}{(\alpha-\beta) U_{k} V_{k}}
\end{aligned}
$$

For the proof, the reader will need the following lemmas.
Lemma 2':

$$
\begin{aligned}
& 2 \sum_{n=1}^{\infty} \frac{1}{\alpha^{k n} U_{k n}}=\frac{1}{\alpha^{k} U_{k}}+U_{k} \sum_{n=1}^{\infty} \frac{1}{U_{k n} U_{k(n+1)}} \\
& 2 \sum_{n=1}^{\infty} \frac{1}{\alpha^{k n} V_{k n}}=\frac{1}{\alpha^{k} V_{k}}+(\alpha-\beta) U_{k} \sum_{n=1}^{\infty} \frac{1}{V_{k n} V_{k(n+1)}}
\end{aligned}
$$

1990]

Lemma 3':

$$
\begin{aligned}
& \sum_{n=1}^{\infty} \frac{1}{\alpha^{k n} U_{k n}}=(\alpha-\beta)\left[L\left(\beta^{2 k}\right)-2 L\left(\beta^{4 k}\right)+2 L\left(\beta^{8 k}\right)\right] \\
& \sum_{n=1}^{\infty} \frac{1}{\alpha^{k n} V_{k n}}=L\left(\beta^{2 k}\right)-2 L\left(\beta^{8 k}\right)
\end{aligned}
$$

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# Applications of Fibonacci Numbers 

## Volume 3

## New Publication

Proceedings of 'The Third International Conference on Fibonacci Numbers and Their Applications, Pisa, Italy, July 25-29, 1988.<br>edited by G.E. Bergum, A.N. Philippou and A.F. Horadam

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## THE DISTRIBUTION OF RESIDUES OF TWO-TERM RECURRENCE SEQUENCES

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(Submitted July 1988)

$$
\begin{aligned}
& \text { Let } U_{0}, U_{1}, A, B \text { be integers and define, for } n \geq 2 \text {, } \\
& U_{n}=A U_{n-1}+B U_{n-2} .
\end{aligned}
$$

For an integer $m>1$, the sequence $\left(U_{n}\right)$ considered modulo $m$ is eventually periodic. We say $\left(U_{n}\right)$ is uniformly distributed modulo $m$ [notation: u.d. (mod $m$ )] if every residue modulo $m$ occurs with the same frequency in any period. In this case, it is clear that the length of any period will be a multiple of $m$. Conditions that $\left(U_{n}\right)$ be u.d. (mod $m$ ) can be found in [2, Theorem A]. Suppose $\left(U_{n}\right)$ is u.d. (mod $\left.p^{k}\right)$ where $p$ is a prime and $k>0$. Let $M \geq 2$ be any integer. We study the relationship between the distribution of $U_{n}$ (mod $M$ ) and $U_{n}$ (mod $\left.M \cdot p^{k}\right)$. For integers $N \geq 2$ and $0 \leq c<N$, denote by $v(N, c)$ the number of times that $c$ occurs as a residue in one shortest period of $U_{n}(\bmod N)$. Our main result can now be stated.

Theorem: Let $p$ be a prime and $k>0$ be an integer such that $U_{n}$ is $u . d .(m o d$ $p^{k}$ ). Say $U_{n}$ has shortest period of length $p^{k} f$ modulo $p^{k}$. Let $M \geq 2$, and assume that $U_{n}$ is purely periodic modulo $M$, with shortest period of length $Q$. Assume $p \nmid Q$. Then, for any $0 \leq a<M$, and $0 \leq b<M \cdot p^{k}$ with $b \equiv a(\bmod M)$,

$$
v\left(M \cdot p^{k}, b\right)=\frac{f}{(Q, f)} \cdot v(M, a) .
$$

We remark that ( , ) denotes the GCD. Also, observe that the hypothesis $p \not Q Q$ yields $P \nmid M$. To prove the Theorem, we make use of a recent result of Vélez [2], which we state here for the reader's convenience.
Lemma: Suppose that $U_{n}$ is u.d. (mod $p^{k}$ ) with shortest period of length $p^{k} f$. Then, for any integer $s \geq 0$, the sequence $U_{s+q f}, q=0,1, \ldots, p^{k}-1$, consists of a complete residue system modulo $p^{k}$.
Proof of Theorem: Let $0 \leq \alpha<M$ and let $v(M, \alpha)=d$. As the Theorem is vacuous if $d=0$, assume $d \geq 1$. Let $w_{1}, w_{2}, \ldots, w_{d}$ be all of the integers $0 \leq$ $w_{i}<Q$ such that $U_{w_{i}} \equiv a(\bmod M)$. Let $0 \leq b<M \cdot p^{k}$, say $b \equiv r\left(\bmod p^{k}\right)$ with $0 \leq r<p^{k}$. Assume $b \equiv a(\bmod M)$. Note that $U_{n}$ has period length

$$
\operatorname{LCM}\left(Q, f p^{k}\right)=\frac{f}{(Q, f)} \cdot Q \cdot p^{k} \text { modulo } M \cdot p^{k} .
$$

For ease of notation, we set $z=f /(Q, f)$. As $\left(M, p^{k}\right)=1$, it suffices, by the Chinese Remainder Theorem, to show that the system

$$
U_{n} \equiv\left\{\begin{array}{l}
a(\bmod M)  \tag{1}\\
p\left(\bmod p^{k}\right)
\end{array}\right.
$$

has exactly $z \cdot d$ solutions, $0 \leq n<z \cdot Q \cdot p^{k}$.
We begin by producing, for each $w_{i}$, solutions $v_{i l}, v_{i 2}, \ldots, v_{i z}$ of the system. Fix $i$. Then

$$
U_{w_{i}}+e Q \equiv a(\bmod M) \text { for all } 0 \leq e<z \cdot p^{k}-1
$$

Let $0 \leq s_{i 1}<s_{i 2}<\cdots<s_{i z}<f$ be all of the distinct integers such that

$$
w_{i} \equiv s_{i 1} \equiv s_{i 2} \equiv \cdots \equiv s_{i z}(\bmod (Q, f)) .
$$

By Vélez's lemma, there exist integers $0 \leq q_{i 1}, q_{i 2}, \ldots, q_{i z} \leq p^{k}-1$ such that $U_{s_{i j}+q_{i j}} \equiv r\left(\bmod p^{k}\right)$, for all $j$.
Then, also, for any $0 \leq t \leq Q /(Q, f)-1$, we have

$$
U_{s_{i j}}+\left(q_{i, j}+t p^{k}\right) f \equiv r\left(\bmod p^{k}\right)
$$

The bounds on $e, t$ guarantee that these subscripts are less than $z \cdot Q \cdot p^{k}$. For each $i, j$, we seek $e=e_{i j}, t=t_{i j}$ in these bounds such that

$$
w_{i}+e_{i j} Q=s_{i j}+\left(q_{i j}+t_{i j} p^{k}\right) f
$$



$$
t \cdot z \cdot p^{k} \equiv-\left(m_{i j}+q_{i j} z\right)\left(\bmod \frac{Q}{(Q, f)}\right)
$$

has a unique solution $t=t_{i j}$ with $0 \leq t_{i j}<\frac{Q}{(Q, f)}-1$. But then

$$
Q \mid(Q, f)\left(m_{i j}+q_{i j} z+t_{i j} \cdot z \cdot p^{k}\right) ;
$$

thus, since $\left(Q, z \cdot Q \cdot p^{k}\right)=Q$, the linear congruence

$$
e Q \equiv(Q, f)\left(m_{i j}+q_{i j} z+t_{i j} \cdot z \cdot p^{k}\right)\left(\bmod z \cdot Q \cdot p^{k}\right)
$$

has $Q$ solutions $0 \leq e<z \cdot q \cdot p^{k}$. Hence, this congruence has a unique solution $e=e_{i j}$ satisfying $0 \leq e_{i j} \leq z \cdot p^{k}-1$. With these values of $e_{i j}, t_{i j}$, we have

$$
w_{i}+e_{i j} Q \equiv s_{i j}+\left(q_{i j}+t_{i j} p^{k}\right) f\left(\bmod z \cdot Q \cdot p^{k}\right),
$$

so equality holds, since both sides are less than $z \cdot Q \cdot p^{k}$. Set $v_{i j}=w_{i}+e_{i j} Q$ for all $i, j$. Then $0 \leq v_{i j}<z \cdot Q \cdot p^{k}$, and each $v_{i j}$ is a subscript that satisfies the system (1), that is, $U_{v_{i j}} \equiv b\left(\bmod M \cdot p^{k}\right)$ for all $i, j$. We claim that the $v_{i j}$ are distinct.

Suppose that $v_{i j}=v_{g h}$. Then $w_{i}+e_{i j} Q=\omega_{g}+e_{g h} Q$ implies $Q \mid\left(w_{i}-w_{g}\right)$. As $0 \leq \omega_{i}, \omega_{g}<Q$, this gives $\omega_{i}=\omega_{g}$, so that $i=g$. Then

$$
s_{i j}+\left(q_{i j}+t_{i j} p^{k}\right) f=s_{i h}+\left(q_{i h}+t_{i h} p^{k}\right) f,
$$

so that $f \mid\left(s_{i j}-s_{i h}\right)$. As $0 \leq s_{i j}, s_{i h}<f$, we have that $s_{i j}=s_{i h}$; therefore, $j=h$. Thus, the $v_{i j}$ are distinct. This shows that, for any $0 \leq \alpha<M$ and any $0 \leq b<M \cdot p^{k}, v\left(M \cdot p^{k}, b\right) \geq z \cdot v(M, a)$. The proof is concluded by observing that

$$
\begin{aligned}
z \cdot Q \cdot p^{k} & =\sum_{b=0}^{M \cdot p^{k}-1} v\left(M \cdot p^{k}, b\right)=\sum_{a=0}^{M-1} \sum_{r=0}^{p^{k}-1} v\left(M \cdot p^{k}, b\right), \text { where } b \equiv\left\{\begin{array}{l}
a(\bmod M) \\
p\left(\bmod p^{k}\right)
\end{array}\right. \\
& \geq \sum_{a=0}^{M-1} \sum_{r=0}^{p^{k}-1} z \cdot v(M, a)=z \cdot p^{k} \sum_{a=0}^{M-1} v(M, a)=z \cdot p^{k} \cdot Q .
\end{aligned}
$$

Hence, equality holds throughout, and the Theorem follows.
Example: Let $A=B=1, U_{0}=0, U_{1}=1$ so that $U_{n}$ is the Fibonacci sequence. Then $U_{n}$ is u.d. (mod 5). Take $M=33$. Then $U_{n}$ has period of length $Q=40$ modulo 33, and one computes that $v(33,1)=5$, whereas $\nu(165,1)=3$. This justifies the hypothesis that $P \not Q$. Moreover, in this case, $\nu(33, a)$ assumes 5 values for $0 \leq a<33$, but $v(165, b)$ assumes only 4 values for $0 \leq b<165$.

In fact, our Theorem asserts that $U_{n}$ has the same number of distinct distribution frequencies modulo $M$ and $M \cdot p^{k}$, whenever $M, p$ satisfy the hypotheses of the Theorem [that is, $\nu(M, *)$ and $\nu\left(M \cdot p^{k}\right.$, *) take on the same number of distinct values]. This provides an alternate method of obtaining the results in [1].

[^0]Masaryk University in Brno, Czechoslovakia, is the only university in the country which subscribes to the Fibonacci Quarterly. Unfortunately, their set is not complete. They need volumes 1-9. If anyone would be interested in donating these volumes to Masaryk University please let the editor of this journal know and he will make arrangements.

# GENERALIZATIONS OF THE DUAL ZECKENDORF INTEGER REPRESENTATION THEOREMS-DISCOVERY BY <br> <br> FIBONACCI TREES AND WORD PATTERNS 

 <br> <br> FIBONACCI TREES AND WORD PATTERNS}

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> "What is the use," thought Alice, "of a book without pictures and conversations?.."
> Alice's Adventures in Wonderland,
> -Lewis Carroll

## 1. Introduction

In this paper we show how the two well-known integer representation theorems which are associated with the name of Zeckendorf may be generalized as dual systems by constructing colored tree sequences whose shade sets partition $Z^{+}=\{1,2, \ldots\}$. Many interesting properties of the representations can be observed directly from the tree diagrams, and the proofs of the properties can truly be said to be "evident" or "obvious"; we shall not translate such proofs into other symbolic forms.

The Zeckendorf theorems are about representations of positive integers as sums of distinct elements of given number sequences. The first theorem is in Lekkerkerker [6], and a dual of it is given by Brown [2]. Early papers on properties of the Zeckendorf integer representations are Zeckendorf [12] and Brown [1]. Klarner [5] gives an excellent review of the literature to 1966, and extends many of the theories to that date. In [3] Carlitz et al. (1972) define Fibonacci representations of integers, and study their properties.

In Turner [7] we showed how to construct certain tree sequences and defined their shade sets, which together demonstrated the Zeckendorf representation theorems. In Zulauf \& Turner [13], we showed how the shade sets could be defined in a set-theoretic notation, and proved the Zeckendorf theorems in a concise manner. In Turner [8] and Turner \& Shannon [9] we defined Fibonacci word patterns and used them to study tree shade sets.

Notation and definitions for integer representations
(i) Let $\mathrm{c}=\left\{c_{1}, c_{2}, c_{3}, \ldots\right\}$ be any sequence of distinct real numbers, and 1et $N \in Z^{+}$(i.e., $N$ is a positive integer). We shall be concerned with representations of $N$ of the form
$N=\sum_{i=1}^{n} e_{i} c_{i}$, where $n \geq 1$ and $e_{i} \in\{0,1\}$ for each $i$.
In this paper $c$ will be a strictly increasing sequence of nonnegative integers. Once c is given, the vector $\mathrm{e}=\left\{e_{1}, \ldots, e_{n}\right\}$ determines the representation.
(ii) (as in [4]): The sequence $c$ is complete with respect to the positive integers if and only if every integer $N \in Z^{+}$has a representation of the form (1.1).
(iii) (see [4]): If the number of elements of $c$ used in a representation is as small as possible, the representation is said to be minimal; if it is the largest possible, then the representation is maximal.

## The dual Zeckendorf theorems

We shall use the notation $Z$ and $Z^{*}$ when referring to these theorems and related properties.

Theorem 1.1 (Z; [6], [12]): Every $N \in Z^{+}$has one and only one representation in the form (1.1) with $c=\left\{u_{n}\right\}=\left\{F_{n+1}\right\}=\{1,2,3,5, \ldots$, the Fibonacci sequence, and with the coefficients $e_{i}$ satisfying $e_{n}=1$, and $e_{i} e_{i+1}=0$ if $1 \leq i<n$.

Moreover, these representations are minimal; and for a given value of $n$ there are $F_{n}$ integers having Zeckendorf representations.

Theorem 1.2 (the dual, $\left.Z^{*} ; ~[2]\right): ~ E v e r y ~ N \in Z^{+}$has one and only one representation in the form (1.1) with $c=\left\{u_{n}\right\}, e_{i} \in\{0,1\}, e_{n}=1$, and $e_{i}+e_{i+1} \neq 0$ if $1 \leq i<n$.

Moreover, these representations are maximal; and for a given value of $n$ there are $F_{n+1}$ integers having Zeckendorf representations.

The generalization in this paper:
In Section 2 we show how to construct classes of colored tree sequences whose shade sets exactly cover $Z^{+}$, and hence derive classes of complete sequenzes of integers (those used to color the trees). From these classes we select =wo which are dual in a sense that generalizes the dual conditions $e_{i} e_{i+1}=0$ and $e_{i}+e_{i+1} \neq 0$ used in the Zeckendorf theorems 1.1 and 1.2 , respectively. [hus, we obtain a class of dual integer representation theorems, of which the ?air $\left(Z, Z^{*}\right)$ is the simplest case.

## 2. Colored Trees and Their Shades

## Definitions

(i) A tree is a set of $n$ nodes (or points), and a set of $n-1$ edges (lines joining pairs of the nodes), having no cycles (paths from a node which return to that node).
(ii) If one node in a tree is distinguished, and labelled as a root, we have a rooted tree.
(iii) If real numbers (in this paper integers) are assigned to the nodes of a rooted tree, we have a number tree. We call the numbers colors of the nodes.
(iv) A node, other than the root, which has only one edge attached to it is called a leaf. There is a unique path from the root to any given leaf. The sum of the colors on a root-to-leaf path is called the shade of the path.
(v) The set of shades of all root-to-leaf paths in a rooted tree is the shade set (or shade) of the tree.

Generation of a 3 -parameter class of colored tree sequences
Suppose we are given a coloring sequence of integers, denoted by $c \equiv c_{0}$, $c_{1}, c_{2}, \ldots, c_{n}, \ldots$ and also an initial sequence of $i$ rooted trees denoted by $T_{0}, T_{1}, \ldots, T_{i-1}$, each of whose nodes is colored by a member of $c$.

Then we can continue the tree sequence in the following way.
For the $n^{\text {th }}$ tree, take a $k$-fork (with $k \leq n$ ) and color its root node $n_{n}$. Select an ordered subsequence of the $T_{0}, T_{1}, \ldots, T_{n-1}$, of length $k$ and using consecutive members, and mount them one by one from left to right on the $k$ prongs on the fork. The following diagram makes this construction clear:

## (2.1) <br> $T_{n} \equiv$


with $k<j \leq n+1$.
Any selection of values for the triple ( $i, j, k$ ) will determine a sequence of colored trees, so the construction just defined determines a 3-parameter family of such trees. We shall also allow $j, k$ to be functions of $n$.
Tree sequences with shade sets exactly equal to $Z_{0}^{+}$:
We investigate now the choices of $c$, the triple ( $i, j, k$ ), and the initial trees such that they will lead to a sequence of trees having shade set $Z_{0}^{+}=\{0$, $1,2, \ldots\}$. We shall require this to happen exactly; which is to say that if $Z_{m}$ denotes the shade set of the $m^{\text {th }}$ tree $T_{m}$ of a sequence we shall require

$$
\bigcup_{m=0}^{\infty} Z_{m}=Z_{0}^{+} \quad \text { and } \quad Z_{u} \cap Z_{v}=\emptyset \text { when } u \neq v
$$

## Examples

Before giving a general result, we shall give three examples to illustrate the various concepts introduced above. The first two provide graphical proofs of the dual Zeckendorf theorems; we treated these in [7] and [13]. The third gives an indication of the generalization we are aiming at, and we give the first seven trees of its sequence.
Example 1 (Zeckendorf, Z)


Example 2 (Zeckendorf dual, 2*)

$$
\begin{array}{ll}
\text { Parameter values: } & (i, j, k)=(2,3,2) \\
\text { Color sequence: } & \left\{0, u_{1}, u_{2}, \ldots\right\} .
\end{array}
$$



Example 3 (gap range 1, 2)


For this last example, it may be observed that the sequence of tree shade sets $\left\{Z_{n}\right\}, n=0,1,2, \ldots$ is

$$
\{\{0\},\{1\},\{2\},\{3,4\},\{5,6\},\{7,8,9\},\{10,11,12,13\}, \ldots\} .
$$

It is also seen that the number of leaves in $T_{n}$ is $\epsilon_{n-2}$ after $n=5$. It should be evident that with this color sequence and method of tree construction the shade set sequence will continue through the positive integers. Thus, the shade set covers $Z_{0}^{+}$exactly. So, given any positive integer $N$, it will correspond to the shade of just one root-to-leaf path in one tree of this sequence of colored rooted trees. It is easy to derive a formula to tell which. We shall not do this here, but rather remark upon the fact that the numbers (colors) on that root-to-leaf path constitute a representation for $N$ as a sum of distinct members of $c$. Thus, $c$ is complete for $Z^{+}$. Moreover, because of the parameter values ( $i, j, k$ ) and the construction process, we can state, from the following illustration, that the representation for $N$ has gaps in its constituent colors of either 1 or 2 (i.e., at least a gap of 1 and at most a gap of 2).

Take $N=12$, for example. This occurs in the shade of $T_{7}$. The third root-to-leaf path from the left gives the representation

$$
12=1+4+7=c_{1}+c_{4}+c_{6}
$$

The binary representation (i.e., the vector of e-values) is ( $0,1,0,0,1,0$, 1). The "gaps" referred to above can now be seen as runs of 0 's occurring setween the $1^{\prime}$ 's. All representations from this tree sequence will have a gap of 1 zero or 2 zeros between every pair of adjacent $l^{\prime}$ 's. A graphical "proof" of this is to write the construction rule thus, where the color gap sizes sccurring are indicated on the fork edges:


It is evident that in every tree beyond the third a 2 or a 1 must occur on every edge, and hence only gaps of 2 and 1 occur in all root-to-leaf path sums.

The reader may care to check that similar reasoning applied to the trees of Examples 1 and 2 will verify the dual Zeckendorf theorems, with their gap properties that $e_{i} e_{i+1}=0$ and $e_{i}+e_{i+1} \neq 0$, respectively.

Before going on to define the dual classes which generalize the Zeckendorf theorems, we give an indication of how studies of Fibonacci word patterns [8] occurring on the tree sequences can provide theorems about properties of integer representations. Referring to Example 2, for instance, suppose we wish to investigate the occurrence of integer representations with the Zeckendorf dual properties and which contain $u_{1}=1$ (i.e., also $e_{1}=1$ ). Examining the trees, we see that $u_{1}=1$ occurs only on leaf nodes. It is easy to derive formulas for the number of $u_{1}^{\prime}$ s occurring in tree $T_{n}$ (it is obviously $F_{n}$ ), and for the pattern of the occurrences. Details of the pattern are given in [8]; briefly, the pattern starting with $T_{1}$ is given by the Fibonacci word juxtaposition recurrence formula

$$
W_{n+2}=W_{n} W_{n+1} \text {, with } W_{1}=1 \text { and } W_{2}=01
$$

This gives the pattern (which is the leaf-node color pattern) 1, 01,101 , 01101, ... . The positions of the 1's are the places $1,3,4,6,8,9, \ldots$ and of the 0 's are $2,5,7,10, \ldots$. . These two sequences are the well-known Wythoff number sequences, given by $\{[\alpha n]\}$ and $\left\{\left[\alpha^{2} n\right]\right\}$, respectively, where $\alpha=$ $\frac{1}{2}(1+\sqrt{5})$ and $[x]$ is the greatest integer function ([8] and [11]). Similar analyses lead to similar conclusions for the placings of the other colors in the tree sequence.

To study relative positions and frequencies of occurrences of integer representations from Example 3, it is necessary to solve the third-order recurrence given for $c_{n}$; and to study the corresponding word pattern recurrence $W_{n+3}=W_{n} W_{n+1}$ with initial words $W_{1}=0, W_{2}=1, W_{3}=2$.

## 3. Generalized Dual Zeckendorf Theorems

We choose parameter values for ( $i, j, k$ ) in (2.1) so that two dual treesequence classes are defined. With suitable choices of initial trees, and of coloring sequences, we shall ensure that the first one (designated the GZclass) generates integer representations such that all gaps in the e-vectors have at least $g^{*}$ zeroes; and that the second one (designated the $\left.G Z^{*}-c l a s s\right)$ generates integer representaitions with all gaps having at most $g^{*}$ zeroes. To be precise, we define a gap $g$ to be a run of $g$ zeroes occurring between two successive 1 's in an e-vector. The conditions "at least $g^{*}$ " and "at most $g^{* "}$ on the gap sizes in the e-vector representations are the dual conditions. We note immediately that the GZ-class will contain the sequence of Example 1 , since the conditions $e_{j e j+1}=0$ and $g \geq 1$ are equivalent. Likewise, the $G Z^{*}{ }_{-}$ class contains the sequence of Example 2 , since the conditions $e_{j}+e_{j+1} \neq 0$ and $g \leq 1$ are equivalent.

The following tables give definitions, and the first few color sequences and corresponding tree sequences as examples.

TABLE 1. Definitions

| GZ-class | $G Z^{*}-\mathrm{class}$ |
| :---: | :---: |
| Gap sizes: $g \geq g^{*}$ | $g \leq g^{*}$ |
| Parameter: $(i, j, k)=\left(g^{*}+1, n+1, n-g^{*}\right)\left(\right.$ for $\left.T_{n}\right)$ | $(i, j, k)=\left(g^{*}+1, g^{*}+2, g^{*}+1\right)$ |
| Color sequence: $c_{n+i}=c_{n}+c_{n+i-1}$ | $c_{n+i}=c_{n}+c_{n+1}+\cdots+c_{n+i-1}$ |
| Initial colors: $0,1,2, \ldots, i$ | $0,2^{0}, 2^{1}, 2^{2}, \ldots, 2^{i-1}$ |
| Initial trees: $\left\{{ }_{r} \mid p=0, \ldots, i-1\right\}$ | The first $i+1$ trees of $\left\{T_{n}\right\}$ are given by: <br> with $2 \leq t \leq i$. |
| General solution for c is given in $\$ 4$ | General solution for c is given in $\$ 4$ |

TABLE 2. Example Sequences

| $g^{*}$ | GZ-Class | $G Z^{*}$-Class |
| :---: | :---: | :---: |
| $\begin{aligned} & 1 \\ & 2 \\ & 3 \end{aligned}$ | $\begin{aligned} & 0,1,2 ; 3,5,8,13, \ldots \\ & 0,1,2,3 ; 4,6,9,13, \ldots \\ & 0,1,2,3,4 ; 5,7,10,14, \ldots \end{aligned}$ | $\begin{aligned} & 0,1,2 ; 3,5,8,13, \ldots \\ & 0,1,2,4 ; 7,13,24,44, \ldots \\ & 0,1,2,4,8 ; 15,29,56, \ldots \end{aligned}$ |
|  | Tree Sequences ( $g \geq g^{*}$ ) | Tree Sequences ( $g \leq g^{*}$ ) |
| 1 2 |  |  |

It should be noted that, since the shades of all the tree sequences in both the $G Z-$ and $G Z^{*}$-classes exactly cover $Z_{0}^{+}$, all the color sequences used (with $3_{0}=0$ deleted) are complete for $Z^{+}$.

Within each pair of tree sequences, for each value of $g^{*}$, the root-to-leaf paths give integer representations using distinct colors and with gaps satisfying $g \geq g^{*}$ and $g \leq g^{*}$, respectively. Those on the tree sequences with $g^{*}=1$ are the dual Zeckendorf representations.

As we said in the Introduction, there is hardly a need for formal proofs of the above statements about the integer representation properties. They all follow by induction, using the definitions of the procedures for constructing the colored trees. Study of the general tree diagram tells all! As Alice thought, in Wonderland: "What is the use of a book without pictures... ." However, to demonstrate the reason for the choices of ( $i, j, k$ ) in the two classes, we shall give some details of the proofs. The key property to establish is that the shade sets of the trees in any sequence partition $Z_{0}^{+}$.
Theorem: Each tree-sequence in the two classes defined in Table 1 has shade set exactly equal to $Z_{0}^{+}=\{0,1,2, \ldots\}$.
Proof: We shall use induction, for sequences in each class.
Case (i) Let $T=\left\{T_{n}\right\}$ be a tree-sequence in the GZ-class.
The first $i$ trees in $T$ have shades $0,1, \ldots$ ( $i-1$ ), respectively, by the definitions of initial colors and initial trees given in Table 1 . The $(i+1)^{\text {th }}$ tree is

$$
T_{i}=\emptyset_{i}^{T_{0}}
$$

since $k=n-g=i-(i-1)=1$ (meaning there is a l-fork), and $n-j+1$ $=n-(n+1)+1=0$ (meaning that $T_{0}$ is mounted on it). Here we have used the formulas given in Table 1 for the parameters ( $i, j, k$ ) in the GZ-class.

Thus, $T_{i}$ has shade $0+i=i$, which continues the shade sequence required by the theorem.

We now make the inductive hypothesis that the shade sets continue as for the theorem, up to the last (rightmost) branch of tree $T_{n}$, with $n>i$.

Referring to the construction diagram (2.1), inserting parameters $j=n+1$ and $k=n-g^{*}$, we find that $T_{0}$ is mounted on the first (leftmost) branch of the $k$-fork used to construct $T_{n}$. This is also true for $T_{n+1}$, etc. Hence, the leftmost branch shade of $T_{n+1}$ is $c_{n+1}+0=c_{n+1}$.

Now, the rightmost branch shade of tree $T_{n}$ is $c_{n}+$ (rightmost branch shade of $\left.T_{n-g^{*}-1}\right)$, which, by the inductive hypothesis, equals $c_{n}+\left(c_{n-g^{*}}-1\right)$. Then, since $c_{n}+c_{n-g^{*}}=c_{n}+c_{n-i+1}=c_{n+1}$ (using parameter and color sequence definitions), we have shown that
(1eftmost branch shade of $T_{n+1}$ ) $=\left(\right.$ rightmost branch shade of $\left.T_{n}\right)+1$.
Hence, the shade of $T_{n+1}$ follows on in natural sequence from that of $T_{n}$. This completes the inductive proof.

Case (ii) Let $T=\left\{T_{n}\right\}$ belong to the $G Z^{*}$-class.
We proceed as for Case (i); we shall omit the details showing that the shades of $T_{0}, T_{1}, \ldots, T_{i+1}$ conform to the theorem.

Assume that the shade of the tree sequence $T_{0}, T_{1}, \ldots, T_{n}$, with $n>i+1$, is a sequence $0,1,2, \ldots, r$. We shall show that the first element of the shade of $T_{n+1}$ is $r+1$.

Let us use the notation $L_{n}, R_{n}$ to mean, respectively, the "leftmost branch shade of tree $T_{n}$ " and the "rightmost branch shade of tree $T_{n}$." We have to show that $R_{n}+1=L_{n+1}$. From the construction diagram (2.1), and inserting the parameters for $j, k$ from Table 1 for the $G Z^{*}$-class, we see that

$$
R_{n}=c_{n}+R_{n} \quad\left(=c_{n}+L_{n}-1\right) ;
$$

and

$$
L_{n+1}=c_{n+1}+L_{n-g^{*}}\left(=c_{n}+c_{n-1}+\cdots+c_{n-g^{*}}+L_{n-g^{*}}\right)
$$

Now

$$
L_{n}=c_{n}+L_{n-g^{*}-1}=c_{n}+\left(L_{n-g^{*}}-c_{n-g^{*}-1}\right)
$$

using the fact that, for $n \geq 1$, the cardinal number of the shade set of $T_{n}$ is equal to $c_{n}$ : this is easily established by induction, for the trees in $G Z^{*}$. So we have

$$
\begin{aligned}
R_{n}+1=c_{n}+L_{n} & =2 c_{n}-c_{n-g^{*}-1}+L_{n-g^{*}} \\
& =c_{n}+c_{n-1}+\cdots+c_{n-g^{*}}+L_{n-g^{*}} \\
& =c_{n+1}+L_{n-g^{*}} \\
& =L_{n+1} .
\end{aligned}
$$

The existence of the generalized dual Zeckendorf integer representations now follows immediately. The proof that the gap sizes satisfy conditions $g \geq g^{*}$ or $g \leq g^{*}$ for tree sequences in the GZ-class or $G Z^{*}$-class, respectively, rests on simple observations of the gaps that can occur [see diagram (2.1)] between $c_{n}$ and the root colors of the $k$ trees $T_{n-j+1}, \ldots, T_{n-j+k}$ used in the construction.

The final table gives the dual representations of $N=1,2, \ldots, 10$ for the cases $g^{*}=1$ and $g^{*}=2$. Note that they are, respectively, minimal and maximal representations. (See Table 3 below.)

## 4. Formulas for the Color Sequences in the Two Classes

In Table 1 we gave the initial values and general recurrence equations for the color sequences in the $G Z-$ and $G Z^{*}$-classes. We end the paper by giving general solutions for the equations, which provide formulas for the terms of the sequences in terms of weighted sums of binomial coefficients. We also give geometrical interpretations for these weighted sums: they are related to the elements on certain diagonals of Pascal's triangle. Thus, in a very nice pictorial way, we have linked the generalizations of the dual Zeckendorf
integer representations to generalizations of the Pascal-Lucas theorem which states that sums of the terms on the $45^{\circ}$ upward diagonals of Pascal's triangle are Fibonacci numbers.

TABLE 3. Dual Integer Representations

| GZ-Class |  |  |  |  | GZ*-Class |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $N$ | $g^{*}=1$ |  | $g^{*}=2$ |  | $N$ | $g^{*}=1$ |  |  | $g^{*}=2$ |  |  |
| 1 | 1 | (1) | 1 | (1) | 1 | 1 |  | (1) | 1 |  | (1) |
| 2 | 2 | (01) |  | (01) | 2 | 2 |  | (01) | 2 |  | (01) |
| 3 | 3 | (001) | 3 | (001) | 3 | $1+$ |  | (11) |  | $+2$ | (11) |
| 4 | $1+3$ | (101) | 4 | (0001) | 4 | $1+$ | 3 | (101) | 4 |  | (001) |
| 5 | 5 | (0001) | $1+4$ | (1001) | 5 | $2+$ | 3 | (011) |  | $+4$ | (101) |
| 6 | $1+5$ | (1001) | 6 | (00001) | 6 | $1+$ | $2+3$ | (111) |  | $+4$ | (011) |
| 7 | $2+5$ | (0101) | $1+6$ | (10001) | 7 | $2+$ | 5 | (0101) |  | $+2+4$ | (111) |
| 8 | 8 | (00001) | $2+6$ | (01001) | 8 | $1+$ | $2+5$ | (1101) |  | $+7$ | (1001) |
| 9 | $1+8$ | (10001) | 9 | (000001) | 9 | $1+$ | $3+5$ | (1011) |  | $+7$ | (0101) |
| 10 | $2+8$ | (01001) | $1+9$ | (100001) | 10 | $2+$ | $3+5$ | (0111) |  | $+2+7$ | (1101) |

The recurrence for the GZ-class
We wish to index and refer to the sequences in the $G Z-c l a s s$, as $g^{*}$ ranges over 1, 2, 3, etc. To this end we add a superscript in brackets, to the expression for the $n^{\text {th }}$ term in the $g^{* t h}$ sequence. Thus, $e_{n}^{\left(g^{*}\right)}$ denotes this expression. Since it is typographically clumsy to use $g^{*}$ as the indexing letter, we shall replace $g^{*}$ by $i$ (note that the values for $i$ as used here and subsequently are 1 less than the ones used for the parameter in Table 1).

The recurrence equation for $c_{n}^{(i)}$ (omitting $c_{0}^{(i)}=0$ from each sequence) is (4.1) $c_{n+i+1}^{(i)}=c_{n+i}^{(i)}+c_{n}^{(i)}$, for $n \geq 1$,
with initial values
(4.2) $\quad c_{s}^{(i)}=s$ for $s=1,2, \ldots, i+1$.
[Note that for $i=1$ this gives the Fibonacci sequence $\left.=\left\{F_{n+1}\right\}.\right]$
The general solution to (4.1) and (4.2), given $i$, is
(4.3) $\quad c_{n-i+1}^{(i)}=\sum_{n=0}^{[n / i]}\binom{n-r i}{p}$, for $n-i+1=1,2,3, \ldots$,
which can be demonstrated by direct substitution, and making use of the identity

$$
\binom{n+1}{n+1}=\binom{n}{n+1}+\binom{n}{n}
$$

for the binomial coefficients.
[We use the normal convention that $\binom{a}{b} \equiv 0$ when $a<b$.]
If $i$ is small, say $i=1,2,3$, or 4 , then we can use Binet-type formulas to calculate the $c_{n}^{(i)}$ efficiently for any $n$. If $i$ is large, then formula (4.3) above is probably the most efficient way to calculate $\mathcal{C}_{n}^{(i)}$ exactly.

For example, if $i=1000$, then $c_{2200}^{(1000)}$ is

$$
\binom{3199}{0}+\binom{2199}{1}+\binom{1199}{2}+\binom{199}{3},
$$

which equals

$$
1+2199+1199 \times 1198 / 2+199 \times 198 \times 197 / 6=2014100
$$

Thus, to calculate $c_{2200}^{(1000)}=2014100$ from the above formula requires only a few additions and multiplications; whereas to calculate it directly from the recurrence relation (4.1) could require that all the $c_{n}^{(1000)}$ be precomputed for $n=$ 1, 2, ..., 2199. Clearly, formula (4.3) is much quicker.

The solution (4.3) also has a nice geometric interpretation, which we show in the final subsection.

The recurrence for the $G Z^{*}$-class
The recurrence equation for $c_{n}^{*(i)}$ (omitting $c_{0}^{*(i)}=0$ from each sequence) is (4.4) $c_{n+i+1}^{\star(i)}=c_{n}^{\star(i)}+c_{n+1}^{\star(i)}+\cdots+c_{n+1}^{\star(i)}$, for $n \geq 1$,
with initial values

$$
\begin{equation*}
c_{s}^{*(i)}=2^{s-1} \text { for } s=1,2, \ldots, i+1 \tag{4.5}
\end{equation*}
$$

[Note that $i=1$ again gives the Fibonacci sequence $\left\{F_{n+1}\right\}$. ]
By considering $c_{n+i+1}^{*(i)}-c_{n+i}^{*(i)}$, and using (4.4), we find that
(4.6) $\quad c_{n+i+1}^{*(i)}=2 c_{n+i}^{\star(i)}-c_{i-1}^{\star(i)}, n=1,2, \ldots$.

We used this equivalent form of the recurrence equation as a first step in obtaining a general solution. The details of our solution method are lengthy, and will be reported elsewhere. Our solution is given next, as (4.7): it may be checked by insertion into (4.4) or (4.6) and use of elementary algebra and manipulations with binomial coefficients.

The general solution to (4.4) and (4.5), given $i$, is

$$
c_{n}^{*(i)}=u_{n}^{(i+1)}-u_{n-1}^{(i+1)} \text { for } n=1,2, \ldots,
$$

where

$$
\begin{equation*}
u_{n}^{(i)}=\sum_{r=0}^{[n / i]} \frac{2^{n-r i}}{(-2)^{r}}\binom{n-r i}{r} \text { for } n=0,1,2, \ldots \tag{4.7}
\end{equation*}
$$

We show in the next subsection how the solutions for the dual pairs of recurrences in the $G Z^{-}$and $G Z^{*}-c l a s s e s$ are neatly related to elements on the diagonals of increasing slope within Pascal's triangle.

## Geometric interpretation

Consider Pascal's triangle, for the binomial coefficients, drawn as a $45^{\circ}$ triangle rather than the usual equilaterial triangle, thus:


Then $c_{n+1-i}^{(i)}$ is just the sum of all the binomial coefficients on the line of slope $i$ that starts on the left end of the $n^{\text {th }}$ row. In particular, the $n^{\text {th }}$ Fibonacci number is the sum of the numbers on a $45^{\circ}$ line (slope $i=1$ ) starting at the $n^{\text {th }}$ row (these lines are the well-known Lucas diagonals). As an example, for the case $i=3$,

$$
c_{7}^{(3)}=c_{9+1-3}^{(3)}=\binom{9}{0}+\binom{6}{1}+\binom{3}{2}=1+6+3=10
$$

The geometric interpretation also suggests the results $c_{n}^{(0)}=2$ and $c_{n}^{(\infty)}=n$ corresponding to lines of slope 0 (horizontal) and $\infty$ (vertical), respectively.

As an example in the $G Z^{*}-c l a s s$, again taking $i=3$, and with $n=9$, we get

$$
\begin{aligned}
c_{9}^{*(3)} & =u_{9}^{(4)}-u_{8}^{(4)}=\left[2^{9}\binom{9}{0}-\frac{2^{5}}{2}\binom{5}{1}-2^{8}\binom{8}{0}-\frac{2^{4}}{2}\binom{4}{1}\right] \\
& =432-224=208
\end{aligned}
$$

Inspection of Pascal's triangle shows that $u_{n}^{(i)}$ is a weighted sum of the elements on the upward diagonal of slope $i$ which begins at the first element of the $n^{\text {th }}$ row: the weights are powers of 2 as given in (4.7).

Hence, $c_{n}^{*(i)}$ is the difference of weighted sums from the adjacent diagonals beginning on the $n^{\text {th }}$ and $(n-1)^{\text {th }}$ rows.

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# THE FIBONACCI SEQUENCE AND THE TIME COMPLEXITY OF GENERATING THE CONWAY POLYNOMIAL AND RELATED TOPOLOGICAL INVARIANTS 

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## 1. Introduction

In this discussion, only tame embeddings of $S^{1}$ in $S^{3}$ that are oriented will be considered. All knots will be represented as regular projections and any projection is assumed to be regular. The reader is expected to know "big 0 " notation; see [2] for example. Some knowledge of NP-Completeness is also useful.

The algorithm analysis of the complexities is relative to the number of crossings of a given knot projection. Also, the analytical creation of the Conway polynomial is in the Class $P$. This is shown by the presentation of a well-known algorithm used for computing the Alexander polynomial which can be easily suited to generate the Conway polynomial in better than $O\left(n^{3}\right)$ time.

The proof of the Conway algorithm having exponential worst case time complexity is based on showing the existence of $n$ crossing knot projections. Given these particular projections, the Conway algorithm may perform $O\left(((1+\sqrt{5}) / 2)^{n}\right)$ operations on the various knot projections which the algorithm derives in order to calculate the Conway polynomial.

Definition 1.1: The crossing number of a knot $K$ is the minimum number of crossings for any regular projection of the knot $K$.
Definition 1.2: A split $Z i n k$ (see [10], [11]) $L \subseteq S^{3}$ is a link $L=L_{1} \cup L_{2}$ where $L_{1}$ and $L_{2}$ are nonempty sublinks and there exist two open balls, $B_{1}$ and $B_{2}$ in $S^{3}$ such that $B_{1} \cap B_{2}=\emptyset$; and $L_{1} \subseteq B_{1}$ and $L_{2} \subseteq B_{2}$.
Definition 1.3: A tangle (see [4]) is a portion of a knot diagram from which there emerge only four arcs from the four "directions" NE, NW, SE, and SW, and possibly some crossings inside.

Examples:


FIGURE 1
A tangle cannot have any knot arcs passing under or over it. The integer tangles pictured above, denoted by the integers +1 and -1 describe all crossings of any knot or link. The following three operations are the elementary knot crossing operations.
Definitions 1.4:

1. Smoothing-This operation takes an oriented +1 or -1 integer tangle and replaces it with an oriented $\infty$ integer tangle.

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2. Changing-This operation takes an oriented +1 or -1 integer tangle and replaces it with an oriented -1 or +1 integer tangle, respectively.
3. Deleting-This operation takes $a+1$ or -1 integer tangle and replaces it with a 0 tangle.


FIGURE 2
Fundamentally, the Conway algorithm is based on the first two knot operations. The operations of changing and smoothing crossings will be focused on, although smoothing and deleting [14] and changing and deleting will also be discussed.

Conceptually, there can be a tree of projections built during the application of elementary crossing operations. This tree of projections will become the basis of the complexity analysis.

A knot's tree of projections relative to changing and smoothing is built as follows:

If the knot $K$ is the unknot, then the tree of projections is the trivial tree with an unknot projection as the root with no children.

Given a knot $K$ that is not the unknot, the tree of projections is the binary tree with a projection of $K$ as the root. Choose a crossing, call it $X$, change the knot projection $K$ at the crossing $X$ to produce the knot projection $L$. Take the knot projection $K$, and the crossing $X$. Smooth it to produce the link projection $R$. The right child of the root projection $K$ is the smoothed knot projection $R$ and the left child of the projection is the changed projection $L$. This process of smoothing and changing of knot projections and nonsplit link projections is recursively continued always placing changed knot or link projections as left subtrees and smoothed knot or link projections become right subtrees. When a subtree becomes the unknot or a split link it is a leaf having no more children. It may be observed that by changing a crossing one can reach either the unknot or a split link. A link's tree of projections is constructed similarly.
Examples:


FIGURE 3

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Notice that the trefoil and the figure eight knots have identically structured trees of projections. Any combinations of the three elementary operations consistently applied forms a tree of projections in a similar fashion. A tree of projections is assumed to be relative to changing and smoothing unless otherwise noted.

To quote Louis Kauffmann [10], a convenient way to get an unknot from a knot is to perform the following operations on the knot diagram [10, p. 79]:

Choose a point $p$ on the diagram and draw knot so that you first draw an over-crossing line at the first encounter with a crossing, under cross at the second encounter, and continue until you return to $p$.
Performing this to a knot projection and then applying Reidemeister moves can produce the unknot in familiar form.
Definition 1.5: An unknot projection developed in this fashion is designated a descending knot projection [6], [7]. A descending knot projection's mirror image is, of course, an ascending knot projection.

Just as there is a rather straightforward algorithm for creating a descending knot projection there is also a straightforward algorithm for detecting an ascending knot projection.

## Definitions 1.6:

1. The unknotting number of a knot is the minimum number of changes that must be performed to produce the unknot starting from any knot projection. The unknotting number of a knot $K$ is denoted by $u(K)$.
2. The unsmoothing number of a knot is the minimum number of smooths that must be performed to produce the unknot or split link from any knot projection. The unsmoothing number of a knot $K$ is denoted by us ( $K$ ).
3. The deleting number of a knot is the minimum number of deletes that must be performed to the crossings to produce a split link or an unknot from any knot projection. The deleting number of a knot $K$ is denoted del(K).

In 1970, J. H. Conway defined a polynomial, $\nabla_{K}(z)$, with integer coefficients for oriented knots and links. The polynomial can be recursively calculated from a regular projection of a knot or link $K$ by consistently applying the knot crossing operations of changing and smoothing.
Theorem 1.7: Every knot $K$ has a well-defined Conway polynomial.
This is well known and a proof can be found in [11].
Surprisingly, the Conway polynomial is well defined independent of the particular sequence of smoothings and changings used to calculate it. One of the most important facts about Conway's algorithm is that it can be applied to any regular projection of $a \operatorname{knot} K$ and it will produce the same polynomial.

## 2. The Algorithm

Algorithm 2.1: (Conway's algorithm, see [4], [10], and [11])
Given: a projection of a knot $K$
Returning: the Conway polynomial of the knot $K$

1. Choose an orientation of the knot projection $K$.
2. If ( $K$ is the unknot), then:

Conway polynomial of the unknot is $1: \quad \nabla_{0}=1$
3. If ( $K$ is a split link), then:

Conway polynomial of the split link is $0: \nabla_{K}=0$

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4. Otherwise, when $K$ is not the unknot and not a split link, the tree of projections is built according to the following formula:

$$
\nabla_{K}-\nabla_{L}=z \nabla_{M}
$$

for $K$, $L$, and $M$ knot and/or link projections identical in all respects except that they differ at one crossing in the following manner:


## FIGURE 4

One must be sure to preserve the orientation and keep track of the multiple of the indeterminant on the right smoothing edge and the constant multiples on the left changing edges.
5. Recursively repeat steps 2 through 5 with the appropriate smoothed and changed knot and/or link projections until the entire tree of projections is built.
6. Return polynomial $\nabla_{K}$.
7. End.

Applying Algorithm 2.1 to a knot projection methodically generates its tree of projections by performing changes and smoothings to the knot projection until there are only projections of unknots or split links left. These changes and smoothings must be performed by adhering to the following formula,

$$
\nabla_{K}-\nabla_{L}=z \nabla_{M}
$$

where $K$, $L$, and $M$ are identical knot projections in every respect except for the following significant differences at only one crossing:


FIGURE 5
Using this relation recursively starting with any knot projection $K$, the projection eventually becomes resolved into either unknots or split links and, therefore, terminates. Notice that during step 4 of Algorithm 2.1 two new knots are created by changing and smoothing. These new knots may have many possible different projections, but any regular projection will suffice. The Conway polynomial of the unknot is defined to be one $\left(\nabla_{0}=1\right)$ and any split link is defined to be zero ( $\nabla_{\text {split }}$ link $=0$ ). The Conway polynomial of both the Trefoil and Figure Eight knots is $z+1$.

The Conway algorithm terminates when all subtrees have become either unknots or split links.

The unknotting number is a lower bound of the number of crossing changes we must perform in order to construct the unknot from a given knot projection.

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Similarly, the unsmoothing number is a lower bound. The next logical question might be: "Given any particular projection of a knot $K$, how fast can Conway's algorithm resolve the projection of the knot $K$ into unknots and split links?"

Lemma 2.2: At most, $\lfloor(n / 2)\rfloor$ is the unknotting number $u(K)$ of any $n$ crossing knot or link $K$.

Proof: First, taking the case of an alternating and oriented knot or link projection $K$ of $n$ crossings, orient the knot or link and then prepare to traverse it. If the traversal begins at an overcrossing, then set out to build a descending knot or link projection; otherwise, construct an ascending knot or link projection.

Without loss of generality, assume that the descending variety will be constructed. Starting traversal at an overcrossing of the knot or link projection every time a crossing is encountered, if it is the first encounter with the crossing, ensure that it is an overcrossing. If it is the first encounter with a crossing, being passed under, change it to an overcrossing. Otherwise, if it is the second encounter with this crossing, then proceed under the crossing. This will construct the descending knot or link in at most

$$
\lceil(n / 2)\rceil-1 \leq\lfloor(n / 2)\rfloor
$$

changes because it was assumed that the traversal started on an overcrossing.
Now we must contend with the nonalternating knot or link projection. Given a knot or link $K$ with $n$ crossings, if $K$ has more overcrossings than undercrossings, then construct the descending knot projection in less than or equal to $\lfloor(n / 2)\rfloor$ crossing changes; otherwise, construct the ascending knot projection similarly. $\square$
Lemma 2.3: At most, $\lfloor(n / 2)\rfloor$ is the deleting number del(K) of any $n$ crossing knot or link $K$.

The proof is similar to the proof of Lemma 2.2 , and is therefore omitted.
Lemma 2.4: Given an $n$ crossing knot or link $L$, at most $n-1$ is the unsmoothing number us( $K$ ).

Proof: Given a projection of the knot $K$ with $n$ crossings, if all $n$ crossings are smoothed, then there will be only unknots and unlinks remaining. After $n-1$ smooths, the knot $K$ will have, at most, only one crossing left, however many associated unlinks are left. The knot with the one remaining crossing must be the unknot. $\square$

Theorem 2.5: At most, given an $n$ crossing knot, the tree of projections relative to changing and smoothing, and smoothing and deleting can have

$$
O\left\{((1+\sqrt{5}) / 2)^{n}\right\} \text { projection nodes. }
$$

[Note that $(1+\sqrt{5}) / 2=\phi$, the "golden" ratio.)
Proof: The tree of projections relative to changing and smoothing follows directly from Lemma 2.2 and Lemma 2.4. Given a knot projection $K$, build a tree of projections that can be described by the recurrence relation

$$
f(n)=f(n-1)+f(n-2)+1
$$

where $f(p)$ is the number of nodes in any tree of projections of a given knot projection with $n$ crossings and, of course, $f(0)=1, f(1)=1$, and $f(2)=3$.

The tree of projections relative to smoothing and deleting follows identically from Lemmas 2.2 and 2.3.

Theorem 2.5 supplies an upper bound on Conway's algorithm (Algorithm 2.1). This is due to the relation

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$$
\begin{equation*}
\nabla_{K}=\nabla_{L}+z \nabla_{M} \tag{1}
\end{equation*}
$$

upon which Conway based his algorithm. Notice the similarity between (1) and the recurrence relation that defines the Fibonacci sequence.

An implementation of Conway's algorithm may not be bounded strictly by the upper bound established by Theorem 2.5, but any implementation of Conway's algorithm can be made to adhere to this upper bound.

A tree of projections relative to changing and deleting has an upper bound of $O\left(2^{(n / 2)}\right)$ given an $n$ crossing knot. Additionally, a tree of projections relative to smoothing and smoothing, or changing and changing, or deleting and deleting all have exponential upper bounds. In summary, we have the following theorem.

Theorem 2.6: At most, the tree of projections relative to any consistently applied elementary knot operation has as an upper bound an exponential number of nodes.

In Theorems 2.5 and 2.6 we have not yet established the existence of any classes of knots which actually adhere to these bounds, i.e., How tight are these bounds?

## 3. The Complexity

A particular class of knots with specific projections illustrates that there exist knot projections whose trees of projection relative to smoothing and changing, and smoothing and deleting actually satisfy the upper bound of

$$
O\left\{((1+\sqrt{5}) / 2)^{n}\right\}
$$

The class of knots is the $(2, n)$ torus knots and links and the Fibonacci knots [14] $F_{n}=1111 \ldots 1$ ( $n$ 1's in Conway's notation, see [4]). The (2, $n$ ) knots and links will all be assumed to have the projections and orientations described below.


FIGURE 6
Standard projection of the $(2, n)$ knots and links
Throughout the rest of this discussion, the (2, $n$ ) torus knots or links will denote the standard projections of the (2, $n$ ) knot or link. Also, given any standard projection of a $(2, n)$ knot or link, smoothing and changing will produce standard projection knots or links. This is being done because standard projections can be maintained throughout the execution of Conway's algorithm. The Fibonacci Knot Class is defined by Turner in [14], which we follow. Just as in [14] a knot from the Fibonacci Knot Class will be denoted as $F_{n}$. In Conway's notation [4], $F_{n}=111 \ldots 1$ ( $n$ ones). This turns out to be useful in the worst case analysis. Several lemmas are now presented without proof, since they can easily be deduced via examination.
Lemma 3.1: Smoothing the specified projection of a torus knot (2, $n$ ), for $m \geq 1$ and $n=2 m+1$, can produce the link projection (2, $n-1$ ). Additionally, smoothing the link projection (2, $n$ ), for $n=2 m$ and $m \geq 1$, can produce a torus knot projection (2, $n-1$ ). We also point out that changing a torus knot

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projection $(2, n)$, where $n=2 m+1$ and $m \geq 1$, can produce the torus knot projection (2, $n-2$ ). And changing the link projection (2, $n$ ), for $n=2$ and $m \geq 1$, can produce the link projection (2, $n-2$ ).

Lemma 3.1 can be verified by examination of parts of the actual torus knot projections.
Lemma 3.2: (Turner [14]) Given a Fibonacci knot $F_{n}$ deleting and smoothing a crossing can create the following projections $F_{n-1}$ and $F_{n-2}$.

An $n$-trefoil is $n$ trefoils connected in the following fashion:


FIGURE 7

Lemma 3.3: Given an $n$-trefoil, any sequence of elementary knot operations forms an exponential tree of projections.

Lemma 3.3 can be proved by using induction.
Definition 3.4: An AVL tree [1], [2] is a binary tree which has the property that from any given node in the tree the depths of the right and left subtrees differ by at most one.

## Examples:



## FIGURE 8

For brevity, from here the discussion is focused on the elementary operations of changing and smoothing.

The following induction establishes that applying Algorithm 2.1 to a (2, $n$ ) knot projection can produce an AVL tree. The Conway algorithm, given a standard projection of a $(2, n)$ knot or link may maintain standard projections throughout smoothing and changing. It is easily proven that any AVL tree has exponential number of nodes, but an exact result will be found.
Theorem 3.5: A torus knot standard projection (2, $n$ ), for $n=2 m+1$ and $m \geq 1$, will produce a tree of projection with $\Theta\left\{((1+\sqrt{5}) / 2)^{n}\right\}$ projection nodes provided standard projections are maintained throughout the computation of the Conway algorithm. The same is true for a ( $2, n$ ) link projection.
Proof: By Lemma 3.1, standard projections can be maintained throughout the construction of a tree of projections of a (2, $n$ ) knot or link projection. Standard projections will be maintained throughout this proof.
Claim: When Conway's algorithm is applied to a (2, $n$ ) knot or link in standard projection, it produces an AVL tree with the number of nodes described by the recurrence relation

$$
f(n)=f(n-1)+f(n-2)+1
$$

Proof by induction:
Basis. The unknot (2, 1) produces a trivial AVL tree. The link (2, 2) produces an AVL tree with 3 knot projections or nodes. The trefoil (2, 3) produces a tree with 5 nodes.

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Inductive hypothesis. Assume that for some number $q$ and some function $f$ the standard knot projection (2, $m$ ) has an AVL tree of projections with $f(m)$ nodes as long as $m<q$. Additionally, the standard link projection (2, $m-1$ ) has an AVL tree of projections with $f(m-1)$ nodes.

Inductive step. By the inductive hypothesis, the link or knot projection
 (2, $q-2$ ) has a tree of projections with $f(q-2)$ nodes in its AVL tree. Now, by adding the +1 integer tangle appropriately to the link or knot projection ( $2, ~ q-1$ ), the knot or link projection ( $2, q$ ) may be produced. Performing a change to the standard projection of the knot or link (2, $q$ ) can produce the knot or link projection (2, $q-2$ ). Smoothing a crossing of the knot or link ( $2, q$ ) may produce the link or knot ( $2, q-1$ ) projection. Making the projection of the knot or link ( $2, q$ ) the root of the tree of projections with the subtrees created by the knot projections (2, $q-1$ ) and (2, $q-2$ ) can make a tree with a total of $f(q-1)+f(q-2)+1$ projection nodes. This is true because the tree of projections of the link or knot (2, $q-1$ ) is an AVL tree, and the tree of projections of the knot or link (2, $q-2$ ) is also an AVL tree by the inductive hypothesis. The link or knot projection (2, q - 1) has smoothing and changing subtrees consisting of the trees of projection (2, $q-2$ ) and (2, $q-3$ ) , respectively. Since depth $((2, q-2))-\operatorname{depth}((2, q-3)) \leq 1$ where depth(K) is the depth of $K$ 's tree of projections, then depth( $2, q-1$ ) depth $((2, q-2)) \leq 1$. Therefore, the knot or link projection ( 2 , $q$ ) can have an AVL tree of projections with number of nodes described by the recurrence relation

$$
f(q)=f(q-1)+f(q-2)+1
$$

This means that applying Conway's algorithm to the standard projection of (2, $n$ ) and preserving standard projections throughout the calculation of the Conway polynomial will produce an AVL tree of projections with the number of nodes described by the recurrence relation

$$
f(n)=f(n-1)+f(n-2)+1
$$

End of induction.
Next, the exact number of nodes, in closed form, in this tree of projections can easily be derived by solving the nonhomogeneous, constant coefficient difference relation of second order:

$$
f(n)-f(n-1)-f(n-2)=1
$$

Given $f(n)-f(n-1)-f(n-2)=1$, where $n \geq 2$, with the boundary conditions $f(0)=1, f(1)=1$, and $f(2)=3$.

The closed form solution is easily derived using the method of variation of constants:

$$
\begin{aligned}
f(n): \\
\begin{aligned}
f(n)=\frac{1+\sqrt{5}}{2 \sqrt{5}}\left(\frac{1+\sqrt{5}}{2}\right)^{n} & +\frac{-1+\sqrt{5}}{2 \sqrt{5}}\left(\frac{1-\sqrt{5}}{2}\right)^{n}-\frac{1}{\sqrt{5}}\left(\frac{1-\sqrt{5}}{2}\right)^{n} \sum_{i=0}^{n-1}\left(\frac{2}{1-\sqrt{5}}\right)^{i+1} \\
& +\frac{1}{\sqrt{5}}\left(\frac{1+\sqrt{5}}{2}\right)^{n} \sum_{i=0}^{n-1}\left(\frac{2}{1+\sqrt{5}}\right)^{i+1}
\end{aligned}
\end{aligned}
$$

Taking the identity

$$
\begin{equation*}
\sum_{i=1}^{n-2}\left(\frac{2}{1+\sqrt{5}}\right)^{i+1}=\frac{1-\sqrt{5}}{1+\sqrt{5}}\left[\left(\frac{2}{1+\sqrt{5}}\right)^{n-1}-\frac{2}{1+\sqrt{5}}\right] \tag{*}
\end{equation*}
$$

which converges to 0 as $n$ approaches $\infty$, noting $(1-\sqrt{5}) / 2<1,2 /(1+\sqrt{5})<1$, and, for some integer $k$ and some constant $A$ when $n \geq k$,

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$$
f(n) \leq A\left(\frac{1+\sqrt{5}}{2}\right)^{n}
$$

we obtain

$$
f(n)=O\left(\left(\frac{1+\sqrt{5}}{2}\right)^{n}\right)
$$

Now, to show $((1+\sqrt{5}) / 2)=O(f(n))$ : Given equation (*) and the fact that $(1-\sqrt{5}) / 2<1$, then choosing a large enough constant $A$ and some number $k$ when $n \geq k$,

$$
((1+\sqrt{5}) / 2)^{n} \leq A((1+\sqrt{5}) / 2)^{n+1} \text { as } n \rightarrow \infty .
$$

Hence,

$$
f(n)=\theta\left\{((1+\sqrt{5}) / 2)^{n}\right\}
$$

This completes the proof. $\square$
Since the tree of projections for a (2, $n$ ) knot projection whose children remain as standard projections has depth of at least $[(n / 2)\rfloor$, it is quite easy to show $f(n)=O\left(2^{\lfloor(n / 2)\rfloor}\right)$, alternatively.

Noting Theorem 2.5 and Theorem 3.5 and given any $n$ crossing knot $K$, call its tree of projections relative to smoothing and changing $P K$. Then the standard projection of the $(2, n)$ knot can have a tree of projections $P(2, n)$ that is larger than or equal to $P K$.

A similar argument, along with Theorem 2.6, illustrates that the Fibonacci Knot Class forms an upper bound on the size of a tree of projections relative to smoothing and deleting. So we have proved
Theorem 3.6: The $(2, n)$ knots and the Fibonacci knots $F_{n}$ have trees of projections relative to changing and smoothing, and smoothing and deleting, respectively, which can be the largest possible given a knot with $n$ crossings.

It is left to the reader to show that the $n$-trefoil will produce a knot which can have the largest possible tree of projections relative to changing and deleting within a constant.

It has been established that there are knot projections (and link projections, for that matter) whose trees of projections can have an exponential number of nodes relative to the number of crossings in the projections. The tree of projections may be built as needed and taken apart immediately afterward.

Any operation on the nodes of the tree of projections of a knot can be considered the fundamental operation. The operation of checking for an unknot or split link seems to fit the bill best (see Algorithm 2.1). This is because the appearance of the unknot or split link indicates a leaf node in the tree of projections with no children; hence, the algorithm may stop smoothing and changing down that particular branch of the tree.

And now a nice application: Denoting the Conway algorithm by $C$.
Theorem 3.7: $\left.C_{\text {worst }}(n)=O((1+\sqrt{5}) / 2)^{n}\right)$; or the Conway algorithm has exponential worst cast time complexity.

The Conway algorithm, given any knot projection $K$ is invariant since the Conway polynomial of the knot $K$ is well defined given any knot projection of $K$ (Theorem 1.7). Let $n$ denote the number of crossings of a particular knot. By Theorem 3.5, using $C$ in constructing the Conway polynomial of the standard projection of the torus knot (2, $n$ ), it is possible to produce an exponential tree of projections. So, $C_{\text {worst }}(n)$ has exponential time complexity dominated by $O(((1+\sqrt{5}) / 2))$ unknotting checks.
Corollary 3.8: Any polynomial invariant constructed by any consistent combination of knot operations (changing, smoothing, and deleting) on a knot

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projection is of exponential worst case cost. Provided the operations are performed to the given knot projection until all of the derived knot or link projections are resolved into unknots and split links at least.

This is true, since these operations are irrespective of the unknown's and their coefficients.

For example, take the Homfly polynomial [6] of a knot $K$, $\operatorname{Homfly}_{K}(x, y, z)$, which can be calculated in a similar fashion to Algorithm 2.1. The Conway polynomial of the same knot $\nabla_{K}(z)$ can then be created by setting $x=1$ and $y=$ 1 , resulting in Homfly $(1,1, z)=\nabla_{K}(z)$.

It might be noted that Conway's algorithm has a constant best case time complexity. It seems to be a very hard problem to decide the average case time complexity of Conway's algorithm.

Theorem 3.9: The act of creating the Conway polynomial is in the class $P$.
This will be proved by the presentation of an algorithm which can determine the Conway polynomial in better than $O\left(n^{3}\right)$ time.

Algorithm 3.10: (A version of Alexander's Algorithm, [3]; see also [13])
Given: a projection of a knot $K$.
Returning: the Conway polynomial of the knot $K$.

1. Choose an orientation of the knot projection $K$. Label the crossings from $X_{1}$ to $X_{n}$ for a knot with $n$ crossings.
2. Create an $n$ by $n$ matrix, calling it mat, filling all entries of mat with zeros.
3. Fill the entries of the matrix as follows:

Each crossing is associated with a column of the matrix.
for col: $=1$ to $n$ do Say crossing col is:


FIGURE 9
Then let

$$
\begin{array}{ll}
\operatorname{mat}[k, & \operatorname{col}]:=1-z \\
\operatorname{mat}[i, & \operatorname{col}]:=z \\
\operatorname{mat}[j, & \operatorname{col}]:=-1
\end{array}
$$

endfor
4. Disregard any row and column of the matrix mat producing a $(n-1) \times(n-1)$ matrix.
5. Calculate the determinant of the matrix produced in step 4.
6. The polynomial created by computing the determinant is the Alexander polynomial.
7. The Alexander polynomial is converted to the Conway polynomial by noting $\Delta_{K}\left(z^{2}\right)=\nabla_{K}\left(z-z^{-1}\right)$ where equality is up to multiples of $\pm z^{n}$.
8. Normalize $\Delta_{K}\left(z^{2}\right)$ into $\nabla_{K}(z)$. Return $\nabla_{K}(z)$.
9. End.

This algorithm clearly terminates, and the computationally time consuming part of Algorithm 3.10 is step 5, calculating the determinant. Step 5 can be done by a straightforward algorithm in $O\left(n^{3}\right)$ time. For example, performing

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Gaussian elimination followed by multiplication along the diagonal computes a determinant in $O\left(n^{3}\right)$ time. There are other algorithms which can compute this in slightly better time. The fact that the matrix is sparse, only having three nonzero elements in each column, can also be taken into account.

This $O\left(n^{3}\right)$ complexity is the same given any $n$ crossing knot projection. Therefore, denoting Alexander's algorithm as Alex, it must be that

$$
\operatorname{Alex}_{\text {best }}(n)=\operatorname{Alex}_{\text {worst }}(n)=\operatorname{Alex}_{\text {average }}(n)=O\left(n^{3}\right)
$$

assuming Gaussian elimination followed by multiplication along the diagonal is used to calculate the determinant. Hence, the act of creating the Conway polynomial is in the class $P$.

## 4. Conclusion

In this paper it is established that the consistent application of elementary knot operations may lead to an exponential number of derived knot projections. This illustrates that Conway's algorithm has exponential worst case time complexity. Moreover, the nonvacuous upper bound on the worst case complexity is based on the golden ratio. It was then illustrated that a determinant based algorithm given by Alexander is of polynomial time complexity; hence, calculating the Conway polynomial is in the class $P$. It is interesting to note that Jaeger [8] has shown the calculation of the Homfly polynomial to be in the class $N P$-Hard.

Corollary 3.8 shows that this complexity analysis can be applied to the calculations of the Jones polynomial [9], the Homfly polynomial [6], the Kauffman polynomial [12, Appendix], and many other polynomial invariants of knots and links.

Hopefully, algorithms for the calculation of knot polynomials will receive more attention in the future. Many interesting questions remain open. It is presently unknown whether any knot polynomial can detect knottedness. Yet Fellows \& Langston [5] have recently nonconstructively shown that detecting knotlessness is in $P$. So the quest is on to find some invariant, knot polynomial or otherwise, that can recognize the unknot in polynomial time.

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# AN ALGEBRAIC IDENTITY AND SOME PARTIAL CONVOLUTIONS 

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Let $\left\{a_{i}\right\}_{i \geq 0}$ and $\left\{b_{i}\right\}_{i \geq 0}$ be any two real or complex number sequences satisfying $a_{i} \neq 0$ and $b_{i} \neq 0$ for $i>0$. Assume that $x, y$, and $z$ are three formal variables. For any natural number $n$, define a formal binomial coefficient as follows:

$$
\begin{equation*}
\binom{x}{n}_{(a)}=\prod_{k=1}^{n} \frac{x-a_{k-1}}{a_{k}}, \text { where }\binom{x}{0}_{(a)}=1 \tag{1}
\end{equation*}
$$

It is obvious that when $a_{k}=k(k=0,1, \ldots),\binom{x}{n}_{(a)}$ reduces to the ordinary binomial coefficient. If we replace $x$ by $1-q^{x}$ and $\alpha_{k}$ by $1-q^{k}(k=0,1$, ...) instead, then $\binom{x}{n}_{(a)}$ becomes the Gaussian binomial coefficient

$$
q^{\binom{n}{2}}\left[\begin{array}{l}
x \\
n
\end{array}\right] .
$$

Based on these preliminaries, we are ready to state our main result.
Theorem: Let $0 \leq m \leq n \leq r$ be three natural numbers. Then the following algebraic identity holds:

$$
\begin{align*}
& \sum_{k=m}^{n}\left\{b_{r-k}\left(x-a_{k}\right) z-\alpha_{k}\left(y-b_{r-k}\right)\right\}\binom{x}{k}_{(a)}\binom{y}{r-k}_{(b)} z^{k}  \tag{2}\\
& =\alpha_{n+1} b_{r-n}\binom{x}{n+1}_{(a)}\binom{y}{r-n}_{(b)^{z^{n+1}}}-\alpha_{m} b_{r-m+1}\binom{x}{m}_{(a)}\left(r-m+1 \begin{array}{c}
y \\
r
\end{array}\right)_{(b)^{z^{m}}} .
\end{align*}
$$

This identity follows from splitting the summand

$$
\begin{aligned}
& \left\{b_{r-k}\left(x-a_{k}\right) z-a_{k}\left(y-b_{r-k}\right)\right\}\binom{x}{k}_{(\alpha)}\binom{y}{r-k}_{(b)^{z}} z^{k} \\
& =a_{k+1} b_{r-k}\binom{x}{k+1}_{(a)}(r-k)_{(b)} z^{k+1}-a_{k} b_{r-k+1}\binom{x}{k}_{(\alpha)}(r-y+1)_{(b)} z^{k}
\end{aligned}
$$

and diagonal cancellation.
Taking $z=1$, (2) reduces to the following.
Corollary: Let $0 \leq m \leq n \leq r$ be three natural numbers, then

$$
\left.\begin{array}{l}
\sum_{k=m}^{n}\left(b_{r-k} x-a_{k} y\right)\binom{x}{k}_{(a)}\binom{y}{r-k}_{(b)}  \tag{3}\\
=a_{n+1} b_{r-n}\binom{x}{n+1}_{(a)}\binom{y}{r}_{(b)}-a_{m} b_{r-m+1}\binom{x}{m}_{(a)}(r-m+1
\end{array}\right)_{(b)} .
$$

For the remainder of the paper, we shall discuss the applications of (2) and (3) to combinatorial identities.

First, letting $m=0, y=m-x$, and $a_{k}=b_{k}=k$ in (3) gives

$$
\begin{equation*}
\sum_{k=0}^{n} \frac{r x-m k}{r x}\binom{x}{k}\binom{m-x}{r-k}=\frac{r-n}{r}\binom{x-1}{n}\binom{m-x}{r-n} . \tag{4}
\end{equation*}
$$

If we define a partial convolution by

$$
\begin{equation*}
S_{m}(x, r, n)=\sum_{k=0}^{n}\binom{x}{k}\binom{m-x}{r-k}, \tag{5}
\end{equation*}
$$

then (4) generates the following recurrence:

$$
S_{m}(x, r, n)=\frac{r-n}{r}\binom{x-1}{n}\binom{m-x}{r-n}+\frac{m}{r} S_{m-1}(x-1, r-1, n-1)
$$

Performing iteration on this recurrence and noting that the closed form of $S_{0}(x, r, n)$ from (4) and (5) is

$$
\begin{equation*}
\sum_{k=0}^{n}\binom{x}{k}\binom{-x}{r-k}=\frac{r-n}{r}\binom{x-1}{n}\binom{-x}{r-n} \tag{6}
\end{equation*}
$$

we have

$$
\begin{equation*}
S_{m}(x, r, n)=\sum_{k=0}^{m} \frac{r-n}{r-k}\binom{m}{k}\binom{r}{k}^{-1}\binom{x-k-1}{n-k}\binom{m-x}{r-n} \tag{7}
\end{equation*}
$$

which contains the following interesting example (cf. Anderson [1]):

$$
\begin{equation*}
\sum_{k=0}^{n}\binom{x}{k}\binom{1-x}{r-k}=\frac{(1-x)(r-1)-n}{r(r-1)}\binom{x-1}{n}\binom{-x}{r-n-1} \tag{8}
\end{equation*}
$$

This identity and (6) are the main results of [1] established by the induction principle.

Rewriting (7) in the form

$$
S_{m}(x, r, n)=(-1)^{m+n} \frac{r-n}{r}\binom{m-x}{r-n}\binom{r-1}{m}^{-1} \sum_{k=0}^{m}\binom{m-r}{m-k}\binom{n-x}{n-k}
$$

and making some trivial modifications, it may be reformulated as

$$
\begin{equation*}
\sum_{i=0}^{n}\binom{n}{i}^{-1}\binom{x}{i}\binom{m-x-r+n}{n-i}\binom{r}{i}=\frac{\binom{n-r}{n}}{\binom{m-r}{m}} \sum_{k=0}^{n}\binom{m-r}{m-k}\binom{n-x}{n-k} \tag{9}
\end{equation*}
$$

Since (9) is a polynomial identity in $r$, it is also true if we replace $r$ by a continuous variable $y$ which provides an algebraic identity. The particular case of $m=0$ in (9) yields the following combinatorial identity:

$$
\begin{equation*}
\sum_{k=0}^{n}\binom{x}{k}\binom{y}{k}\binom{n}{k}^{-1}\binom{n-x-y}{n-k}=\binom{n-x}{n}\binom{n-y}{n} \tag{10}
\end{equation*}
$$

Next, taking $a_{i}=b_{i}=i$ and replacing $r$ and $n$ by $m+n$ in (3), we have

$$
\begin{equation*}
\sum_{k=0}^{n}\{(m+k) y-(n-k) x\}\binom{x}{m+k}\binom{x-1}{n-k}=m y\binom{x}{m}\binom{y-1}{n} \tag{11}
\end{equation*}
$$

Putting $x=y$ and $m=n+1$ in (11), we obtain the following identity,

$$
\begin{equation*}
\sum_{k=0}^{n} \frac{2 k+1}{n+k+1}\binom{x}{n-k}\binom{x-1}{n+k}=\binom{x-1}{n}^{2} \tag{12}
\end{equation*}
$$

which reduces to an identity of Riordan ([3], p. 18) for $x=n-m$.
If we let $m \rightarrow m+s, x \rightarrow 2 m+s$, and $y \rightarrow 2 n+s$, alternatively, then (11) degenerates to Prodinger's generalization for Riordan's identity (cf. [2], and [3], p. 89):

$$
\begin{equation*}
\sum_{k \geq 0}(2 k+s)\binom{2 m+s}{m-k}\binom{2 n+s}{n-k}=\frac{(m+s)(n+s)}{m+n+s}\binom{2 m+s}{m}\binom{2 n+s}{n} \tag{13}
\end{equation*}
$$

Finally, letting $x=y=1-q^{t}, \alpha_{i}=b_{i}=1-q^{i}$, and $m=0$ in (3), we obtain the following q-binomial convolution formula by simple computation:

$$
\sum_{k=0}^{n} \frac{1-q^{r-2 k}}{1-q^{r-n}}\left[\begin{array}{l}
t  \tag{14}\\
k
\end{array}\right]\left[\begin{array}{c}
t \\
r-k
\end{array}\right] q^{(n-k)(r-n-k-1)}=\left[\begin{array}{c}
t-1 \\
n
\end{array}\right]\left[\begin{array}{c}
t \\
r-n
\end{array}\right]
$$

When $q \rightarrow 1$, (14) reduces to the ordinary binomial identity:

$$
\begin{equation*}
\sum_{k=0}^{n} \frac{r-2 k}{r-n}\binom{x}{k}\binom{x}{r-k}=\binom{x-1}{n}\binom{x}{r-n} \tag{15}
\end{equation*}
$$

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# ON CERTAIN NUMBER-THEORETIC INEQUALITIES 

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## 1. Introduction

This note deals with certain inequalities involving the elementary arithmetic functions $d(r), \phi(r)$, and $\sigma_{k}(r)$ and their unitary analogues. We recall that a divisor $d$ of $r$ is called unitary [2] if $(d, r / d)=1$. Let $e \equiv 1$, $I_{k}(r)$ $=r k(k \geq 0)$, and $\mu$ denote the Möbius function. In terms of Dirichlet convolution, denoted by (•), we have [1]:

$$
\left.\begin{array}{rl}
d(r) & =(e \cdot e)(r) \\
\phi(r) & =(I \cdot \mu)(r) \\
\sigma_{k}(r) & =\left(I_{k} \cdot e\right)(r)
\end{array}\right\} \text { where } I(r)=r .
$$

The unitary convolution of two arithmetic functions $f$ and $g$ is defined by
(1.1) $(f \oplus g)(r)=\sum_{d \|_{r}} f(d) g\left(\frac{r}{d}\right)$,
where $d \| r$ means that $d$ runs through the unitary divisors of $r$. The unitary analogue $\mu^{*}$ of $\mu$ is given by [2]
(1.2) $\mu^{*}(r)=(-1)^{\omega(r)}$,
where $\omega(r)$ denotes the number of distinct prime factors of $r$ with $\omega(1)=0$. The unitary analogue $\phi^{*}$ [2] of the Euler totient is given by
(1.3) $\phi^{*}(r)=\left(I \oplus \mu^{*}\right)(r)$.

The unitary analogues of $d$ and $\sigma_{k}$ are $d^{*}$ and $\sigma_{k}^{*}$ and
(1.4) $d^{*}(r)=2^{\omega(r)}$,
$\omega(r)$ being as defined in (1.2);

$$
\sigma_{k}^{\star}(r)=\left(I_{k} \oplus e\right)(r)
$$

For properties of $\sigma_{k}^{*}$, see [5]. It is known that $d^{*}, \phi^{*}$, and $\sigma_{k}^{*}$ are multiplicative functions. Further, given a prime $p, m \geq 1$,

$$
\left\{\begin{array}{l}
d^{*}\left(p^{m}\right)=2  \tag{1.5}\\
\phi^{*}\left(p^{m}\right)=p^{m}-1 \\
\sigma_{k}^{*}\left(p^{m}\right)=p^{m k}+1
\end{array}\right.
$$

Let $\phi_{k}=\left(I_{k} \cdot \mu\right) . \phi_{k}(r)$ is multiplicative in $r$.
From the structure of $\phi_{k}$ and $\sigma_{k}$, we note that

$$
\begin{aligned}
\left(\phi_{k} \cdot \sigma_{k}\right) & =\left(I_{k} \cdot \mu\right) \cdot\left(I_{k} \cdot e\right) \\
& =\left(I_{k} \cdot I_{k}\right) \cdot(\mu \cdot e) \\
& =\left(I_{k} \cdot I_{k}\right) \text { as } \mu \text { is the Dirichlet inverse of } e .
\end{aligned}
$$

or
(1.6) $\quad \sum_{d \mid r} \phi_{k}(d) \sigma_{k}\left(\frac{r}{d}\right)=r^{k} d(r) \quad(k \geq 1)$.

It follows that
$\phi_{k}(r)+\sigma_{k}(r)=\sum_{\substack{d \mid r \\ d \neq 1, d \neq r}} \phi_{k}(d) \sigma_{k}\left(\frac{r}{d}\right)=r^{k} d(r)$.
Therefore,
(1.7) $\quad \phi_{k}(r)+\sigma_{k}(r) \leq r^{k} d(r)$
with equality if and only if $r$ is a prime.
In arriving at (1.7), we have used the fact that $\phi_{k}$ and $\sigma_{k}$ assume only positive values.

Defining $\phi_{k}^{*}=I_{k} \oplus \mu^{*}$, and noting that
(1.8) $\quad \phi_{k}^{*} \oplus \sigma_{k}^{*}=r^{k} d^{*}(r)$,
we have
Theorem 1: $\phi_{k}^{*}(r)+\sigma_{k}^{*}(r) \leq r^{k} d^{*}(r)$ with equality if and only if $r$ is a prime power.

Further, using the fact that

$$
\phi_{\hat{k}}^{*} \oplus d *=\sigma_{\hat{k}}^{*},
$$

we also obtain
Theorem 2: $\phi_{k}^{*}(r)+d^{*}(r) \leq \sigma_{k}^{*}(r)$ with equality if and only if $r$ is a prime power.

We remark that Theorem 2 is analogous to the inequality involving $\phi, d$, and $\sigma$, see [4], [6].

Using the multiplicativity of $\phi_{k}^{*}$ and $\sigma_{k}^{*}$, one could also prove
Theorem 3: For $k \geq 1$,

$$
\frac{1}{\zeta(2 k)}<\frac{\sigma_{k}^{*}(r) \phi_{k}^{*}(r)}{r^{2}}<1
$$

where $\zeta(s)$ is the Riemann $\zeta$-function.
Now, the AM-GM inequality yields
(1.9) $\frac{\sigma_{k}(r)}{d(r)} \geq r^{k / 2}$ (see [9])
and
(1.10) $\frac{\sigma_{k}^{*}(r)}{d^{*}(r)} \geq r^{k / 2}$.

The aim of this note is to establish a few more inequalities which come out as special cases of certain general inequalities found in [3] and [7].

Let

$$
\begin{aligned}
& 0<a \leq a_{i} \leq A \\
& 0<b \leq b_{i} \leq B
\end{aligned} \quad(i=1,2, \ldots, s)
$$

where $a_{i}, b_{i}(i=1,2, \ldots, s), a, A, b, B$ are real numbers. Then, from [7],

$$
\text { (1.11) } \frac{\left(\sum_{i=1}^{s} a_{i}^{2}\right)\left(\sum_{i=1}^{s} b_{i}^{2}\right)}{\left(\sum_{i=1}^{s} a_{i} b_{i}\right)^{2}} \leq \frac{(A B+a b)^{2}}{4 A B a b}
$$

Next, let

$$
0 \leq a_{1}^{(k)} \leq a_{2}^{(k)} \leq \cdots \leq a_{s}^{(k)} \quad(k=1,2, \ldots, m)
$$

Then, an inequality due to Tchebychef [3] states that:
(1.12) $\left(\frac{\sum_{i=1}^{s} \alpha_{i}^{(1)}}{s}\right) \cdots\left(\frac{\sum_{i=1}^{s} \alpha_{i}^{(m)}}{s}\right) \leq \frac{\sum_{i=1}^{s} \alpha_{i}^{(1)} \cdots \alpha_{i}^{(m)}}{s}$

The inequalities derived in Section 2 are essentially illustrations of (1.11) and (1.12).

## 2. Inequalities

Theorem 4: For $k \geq 0$,
(2.1) $\frac{\sigma_{k}(r)}{d(r)} \leq \frac{r^{k}+1}{2}$
and
(2.2) $\quad \frac{\sigma_{k}^{*}(r)}{d^{*}(r)} \leq \frac{r^{k}+1}{2}$.

Proof of (2.1): Let $d_{1}, \ldots, d_{s}$ be the divisors of $r$. We appeal to (1.11) by taking $a_{i}=d_{i}^{k / 2}, b_{i}=d_{i}^{-m / 2}, A=r^{k / 2}, a=1, b=r^{-m / 2}, B=1$. Then

$$
\frac{\sigma_{k}(r) \sigma_{m}(r)}{r^{m}\left(\sigma_{(k-m) / 2}^{(r)}\right)^{2}} \leq \frac{\left(r^{k / 2}+r^{-m / 2}\right)^{2}}{4 r^{k / 2-m / 2}}
$$

or
(2.3) $\quad \frac{\left(\sigma_{k}(r) \sigma_{m}(r)\right)^{1 / 2}}{\sigma_{(k-m) / 2}^{(r)}} \leq \frac{1}{2 r^{(k-m) / 2}}\left(r^{(k+m) / 2}+1\right)$

Setting $m=k$ in (2.3), we obtain (2.1).
Similarly, by considering the unitary divisors of $r$, we arrive at (2.2).
In view of (1.9) and (1.10), we also have
Corollary:
(2.4) $\quad r^{k / 2} \leq \frac{\sigma_{k}(r)}{d(r)} \leq \frac{r^{k}+1}{2}$
and
(2.5) $\quad r^{k / 2} \leq \frac{\sigma_{k}^{*}(r)}{d^{*}(r)} \leq \frac{r^{k}+1}{2}$

Theorem 5: For $k, m \geq 0$,
(2.6) $\frac{\sigma_{k+m}(r)}{\sigma_{m}(r)} \geq r^{k / 2}$
and
(2.7) $\frac{\sigma_{k+m}^{*}(r)}{\sigma_{m}^{*}(r)} \geq r^{k / 2}$.

Proof of (2.6): Let $d_{1}, \ldots, d_{s}$ be the divisors of $r$. We appeal to (1.12) with

$$
a_{i}^{(1)}=d_{i}^{k_{1}}, \ldots, a_{i}^{(m)}=d_{i}^{k_{m}}(i=1,2, \ldots, s)
$$

where $k_{1}, \ldots, k_{m}$ are positive numbers.
Then,

$$
\frac{\sigma_{k_{1}}+\cdots+k_{m}(\Upsilon)}{s} \geq \frac{\sigma_{k_{1}}(r)}{s} \cdots \frac{\sigma_{k_{m}}(r)}{s}
$$

with $s=d(r)$. From (1.9), we obtain
(2.8) $\frac{\sigma_{k_{1}}+\cdots+k_{m}(r)}{\sigma_{k_{i}}(r)} \geq r^{\frac{1}{2} \sum_{j \neq i} k_{j}}$.

Writing $m=2$, we get

$$
\frac{\sigma_{k_{1}+k_{2}}(r)}{\sigma_{k_{2}}(r)} \geq r^{k_{1} / 2},
$$

which proves (2.6).
The proof of (2.7) is similar and is omitted here.
Remark: Inequalities (2.6) and (2.7) generalize (1.9) and (1.10), respectively.
In this connection, we point out that analogous to the inequality $\phi(r) d(r) \geq r[8]$, one could prove using multiplicativity of $\phi_{k}^{*}$ and $d^{*}$ that
Theorem 6: For $k \geq 1$,
(2.9) $d^{*}(r) r^{k} \leq \phi_{k}^{*}(r)\left(d^{*}(r)\right)^{2} \leq r^{2 k}$ 。

The proof of (2.9) is omitted.

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# LENGTH OF THE $n$-NUMBER GAME 

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The $n$-number game is defined as follows. Let $S=\left(s_{1}, s_{2}, \ldots, s_{n}\right)$ be an $n$-tuple of nonnegative integers. A new $n$-tuple $D(S)=\left(\hat{s}_{1}, \hat{s}_{2}, \ldots, \hat{s}_{n}\right)$ is obtained by taking numerical differences; that is, $\hat{s}_{i}=\left|s_{i}-s_{i+1}\right|$. Subscripts are reduced modulo $n$ so that $\hat{s}_{n}=\left|s_{n}-s_{1}\right|$. The sequence $S, D(S)$, $D^{2}(S), \ldots$ is called the n-number game generated by $S$. To see that a game contains only a finite number of distinct tuples let $|S|=\max \left\{s_{i}\right\}$ and observe that $|S| \geq|D(S)|$. Since there are only a finite number of $n$-tuples with entries less than or equal to $|S|$, eventually repetition must take place. When $n=2^{w}$, it is well known that every game terminates with ( $0,0, \ldots, 0$ ). That this is not the case for other values of $n$ is easily seen by considering the following 3-tuple:

$$
\begin{aligned}
R & =(1,0,0) \\
D(R) & =(1,0,1) \\
D^{2}(R) & =(1,1,0) \\
D^{3}(R) & =(0,1,1) \\
D^{4}(R) & =(1,0,1)=D(R)
\end{aligned}
$$

The tuples $D(R), D^{2}(R)$, and $D^{3}(R)$ form what is called a cycle.
For any $n$-tuple $S$, we say the game generated by $S$ has length $\lambda$, denoted by $L(S)$, if $D^{\lambda}(S)$ is in a cycle, but $D^{\lambda-1}(S)$ is not. Thus, in the example above, $L(R)=1$, while $L(D(R))=0$. For each $n$, the length of games is unbounded. That is, for any $\lambda$, there exists an $n$-tuple $S$ such that $L(S)>\lambda$. On the other hand, for tuples $S$ with $|S| \leq m$, there is a game of maximum length, since there are only a finite number of such tuples. We introduce the following notation:

$$
\begin{aligned}
& \mathscr{S}_{n}(m)=\{S \mid S \text { is an } n \text {-tuple with }|S|=m\}, \\
& \mathscr{L}_{n}(m)=\max \left\{L(S) \mid S \in \mathscr{S}_{n}(m)\right\} .
\end{aligned}
$$

On occasion, when the context is clear, we will drop the subscript. The values of $\mathscr{L}_{4}(\mathrm{~m})$ and $\mathscr{L}_{7}(\mathrm{~m})$, along with tuples giving games of maximum lengths, have been determined in [10] and [6]. We consider this question when $n$ is not a power of 2. We first find an upper bound on $\mathscr{L}_{n}(m)$. Then we show that this bound is actually realized when $n=2^{\omega}+1$.

Before proceeding, a few additional comments are in order. Observe that, for any tuple $S$, if we multiply all the entries by a constant $c$ and denote the resulting tuple by $c S$, then
(1) $\quad D(c S)=c D(S)$.

Additionally, if all the nonzero entries of $S$ are equal with $S \in \mathscr{F}(m)$, then $S=m E$ for some $E \in \mathscr{F}(1)$. In particular, an entry $e_{i}$ in $E$ equals 1 if and only if the corresponding entry in $S, s_{i}$, equals $m$.

Since a game concludes when a cycle is reached, it is important to be able to identify those tuples which occur in a cycle. This author did that in [5]. The following theorem gives the salient facts from that work. We say that an $n$-tuple $S$ has a predecessor if $S=D(R)$ for some $n$-tuple $R$.
Theorem 1: Let $n=k r$ where $k=2^{k}$ and $r$ is odd with $r>1$. Suppose $S$ is an n-tuple. Then
$S$ has a predecessor if and only if there exist values $\varepsilon_{\ell} \in\{-1,1\}$, $\ell=1,2, n$, such that

$$
\sum_{\ell=1}^{n} \varepsilon_{l} s_{l}=0
$$

$$
\begin{equation*}
S \text { is in a cycle if and only if all its entries are } 0 \text { or }|S| \text { and } \tag{ii}
\end{equation*}
$$

$$
\sum_{j=0}^{r-1} e_{i+j k} \equiv 0(\bmod 2), \text { for } i=1, \ldots, k
$$

where $S=|S| \cdot E$ with $E=\left(e_{1}, e_{2}, \ldots, e_{n}\right) \in \mathscr{S}(1)$.
Part (ii) guarantees that when $n$ is not a power of 2 , there are nontrivial cycles; indeed, for $n$ odd, the tuple $\mathbb{E}=(0, \ldots, 0,1,1)$ is in a cycle. Moreover, (ii) along with (1) gives
(2) $L(c S)=L(S)$.

## 2. A Bound on $\mathscr{L}_{n}(m)$

For $S \in \mathscr{S}_{n}(m)$, we say that $S$ has $\mu 0^{\prime} s$ and $m^{\prime} s$ in a row, denoted by $\mu(S)$, if the following conditions are met: there exists an integer $\eta$ such that $s_{i} \in$ $\{0, m\}$ for $i=\eta, \eta+1, \ldots, \eta+\mu-1$, at least one of these $s_{i}$ equals $m$, and $\mu$ is as large as possible. As usual, we reduce subscripts modulo $n$. Thus, for example,

$$
\mu(S)=6 \text { when } S=(3,2,3,0,1,3,0,3,0,0)
$$

Loosely speaking, a tuple $S$ will produce a long game if, at each step, $\mu\left(D^{k}(S)\right)$ is as large as possible. In determining an upper bound on $\mathscr{L}_{n}(m)$, the following lemmas will be useful.
Lemma 1: Let $S \in \mathscr{S}_{n}(m), \mu(S)=t$, and $t<n$. If $D(S) \in \mathscr{C}_{n}(m)$, then $\mu(D(S)) \leq$ $t-1$.

Proof: By hypothesis, for some $n$, we have

$$
\begin{array}{ll}
s_{i} \in\{0, m\} & \text { for } i=n, \eta+1, \ldots, \eta+t-1 \\
s_{i}=m & \text { for some } i, \eta \leq i \leq \eta+t-1 \\
1 \leq s_{n-1}, & s_{\eta+t} \leq m-1
\end{array}
$$

As before, let $D(S)=\left(\hat{s}_{1}, \ldots, \hat{s}_{n}\right)$. Then

$$
\begin{aligned}
& \hat{s}_{i} \in\{0, m\} \text { for } i=\eta, \eta+1, \ldots, \eta+t-2, \\
& 1 \leq \hat{s}_{\eta-1}, \hat{s}_{\eta+t-1} \leq m-1 . \\
& \text { Hence, if }|D(S)|=m \text {, then } \mu(D(S)) \leq t-1 .
\end{aligned}
$$

At first glance, it might seem in Lemma 1 that, if $|D(S)|=m$, then $\mu(D(S))$ must equal $t-1$. It is possible, however, to have strict inequality. This would occur if $\hat{s}_{i}=0$ for $\eta \leq i \leq \eta+t-2$, while $\hat{s}_{j}=m$ for some other $j$.
Lemma 2: Suppose that $S \in \mathscr{S}_{n}(m)$ and not all the nonzero entries equal $m$. Then $\left|D^{n-1}(S)\right| \leq m-1$. Further, if $S$ has a predecessor, then $\left|D^{n-2}(S)\right| \leq m-1$.
Proof: Let $\mu(S)=t$. By hypothesis, $t \leq n-1$, and if $S$ has a predecessor, then by Theorem 1 (i), $t \leq n-2$. In either case, Lemma 1 applies. So, if $\left|D^{i}(S)\right|=m$, for $i=1, \ldots, t-1$, then $\mu\left(D^{i}(S)\right) \leq t-i$. of course, if $\mu\left(D^{j}(S)\right)=1$, then $\left|D^{j+1}(S)\right| \leq m-1$. Thus, $\left|D^{t}(S)\right| \leq m-1$.

In a moment we will consider those tuples in which all nonzero entries equal $m$. In that case, $S=m E$ for some $E \in \mathscr{S}(1)$. For tuples in $\mathscr{S}_{n}(1)$, the following is useful. Let $\mathbb{A}=\mathbb{Z}_{2}[t] / \mathscr{I}$ where $\mathbb{Z}_{2}[t]$ is the polynomial ring over
$\mathbb{Z}_{2}$ and $\mathscr{I}$ is the principal ideal generated by $t^{n}+1$. We associate with $E=$ $\left(e_{1}, \ldots, e_{n}\right) \in \mathscr{S}_{n}(1)$, the polynomial

$$
\mathscr{P}_{E}(t)=e_{n}+e_{n-1} t+\cdots+e_{2} t^{n-2}+e_{1} t^{n-1} \text { in } \mathbb{A}
$$

Since $\hat{e}_{i}=\left|e_{i}-e_{i+1}\right|=e_{i}+e_{i+1}$ in $\mathbb{Z}_{2}$ and $t^{n}=1$ in $\mathbb{A}$,

$$
\begin{align*}
\mathscr{P}_{D(E)}(t) & =\left(e_{n}+e_{1}\right)+\left(e_{n-1}+e_{n}\right) t+\cdots+\left(e_{2}+e_{3}\right) t^{n-2}+\left(e_{1}+e_{2}\right) t^{n-1}  \tag{3}\\
& =(1+t) \mathscr{P}_{E}(t)
\end{align*}
$$

Lemma 3: Let $n=k r$, where $k=2^{k}$ and $r$ is odd with $r>1$. Suppose $S \in \mathscr{S}_{n}(m)$ and all the nonzero entries equal $m$. Then $L(S) \leq k$. Further, if $S$ has a predecessor, then $L(S) \leq k-1$.
Proof: As usual, we let $S=m E$, where $E=\left(e_{1}, \ldots, e_{n}\right) \in \mathscr{P}(1)$. For the first part, by (2), we need only show that $D^{k}(E)$ is in a cycle. Using (3), we find

$$
\begin{aligned}
\mathscr{P}_{D^{k}(E)}(t) & =(1+t)^{k} \mathscr{P}_{E}(t) \\
& =\left(1+t^{k}\right) \mathscr{P}_{E}(t) \\
& =\left(1+t^{k}\right)\left(e_{n}+e_{n-1} t+\cdots+e_{2} t^{n-2}+e_{1} t^{n-1}\right) \\
& =\sum_{\ell=0}^{k-1}\left(e_{n-\ell}+e_{k-\ell}\right) t^{\ell}+\sum_{\ell=k}^{n-1}\left(e_{n-\ell}+e_{n+k-\ell}\right) t^{\ell} \text { in } \mathbb{A} .
\end{aligned}
$$

The second equality holds since $k$ is a power of 2 and so all the binomial coefficients in $(1+t)^{k}$ except for the first and last are even. From the above, we see that

$$
D^{k}(E)=\left(e_{1}+e_{k+1}, e_{2}+e_{k+2}, \ldots, e_{n-k}+e_{n}, e_{n-k+1}+e_{1}, \ldots, e_{n}+e_{k}\right) .
$$

We now check condition (ii) of Theorem 1 . In doing so, we use the fact that $n-k=(r-1) k$. For $i=1$, we have

$$
\left(e_{1}+e_{k+1}\right)+\left(e_{k+1}+e_{2 k+1}\right)+\cdots+\left(e_{n-k+1}+e_{1}\right) \equiv 0(\bmod 2)
$$

Similarly, (ii) holds for all other values of $i$. Thus, $D^{k}(E)$ is in a cycle and $L(E) \leq k$.

For the second part, it is also sufficient to show that $L(E) \leq k-1$. Consider the tuple $F=\left(f_{1}, f_{2}, \ldots, f_{n}\right) \in \mathscr{F}(1)$ defined by

$$
f_{1}=0, f_{i}=e_{1}+e_{2}+\cdots+e_{i-1}(\bmod 2), i=2, \ldots, n
$$

Since $S$ has a predecessor, $E$ does as well; because the entries of $E$ are either 0 or 1 , Theorem $1(i)$ gives

$$
e_{1}+e_{2}+\cdots+e_{n} \equiv 0(\bmod 2)
$$

This means that $f_{n}=e_{n}$ and so $D(F)=E$. Thus, $L(E)=L(D(F)) \leq k-1 . \square$
Theorem 2: Let $n=k r$, where $k=2^{k}$ and $r$ is odd with $r>1$. Then $\mathscr{L}_{n}(m) \leq$ $(m-1)(n-2)+k$.
Proof: Let $S \in \mathscr{S}_{n}(m)$. If all the nonzero entries of $S$ are equal, then by Lemma $3, L(S) \leq k$ and so the theorem holds. Otherwise, by Lemma $2,\left|D^{n-1}(S)\right| \leq m-1$. Continuing, suppose that, for some $\ell=1, \ldots, m-2$, all the nonzero entries of $D^{\ell(n-2)+1}(S)$ are equal. Then, again by Lemma 3 , $L\left(D^{\ell(n-2)+1}(S)\right) \leq k-1$, which means $L(S) \leq \ell(n-2)+k$. On the other hand, if the latter condition does not hold, then, by Lemma 2, $\left|D^{(m-1)(n-2)+1}(S)\right| \leq 1$. Another application of Lemma 3 gives the desires result.

If there is a tuple $S \in \mathscr{S}_{n}(m)$ with $L(S)=(m-1)(n-2)+k$, then the proof of Theorem 2 tells us what the tuples in the game must look like.

Corollary 1: Let $n=k r$, where $k=2^{k}$ and $r$ is odd with $r>1$. If

$$
\mathscr{L}_{n}(m)=(m-1)(n-2)+k
$$

then there exists $S \in \mathscr{S}_{n}(m)$ such that
(i) $\left|D^{\ell(n-2)+1}(S)\right|=m-\ell$ and $\mu\left(D^{\ell(n-2)+1}(S)\right)=n-2$ for $\ell=0, \ldots, m-1$,
(ii) $L\left(D^{(m-1)(n-2)+1}(S)\right)=k-1$.

Proof: This follows immediately from the proof of Theorem 2.
In a moment we will state a condition for the existence of a game of maximum length in terms of the $n$-tuple ( $0, \ldots, 0,1,1$ ). Before proceeding, two comments are in order. First, if the entries of an $n$-tuple are rearranged so that adjacent elements remain adjacent, then similar games result. Or, more precisely, if $S=\left(s_{1}, s_{2}, \ldots, s_{n}\right)$ and $\sigma_{1}$ is a permutation contained in the dihedral group $\mathscr{D}_{n}$, then
(4) $\quad D\left(\sigma_{1}(S)\right)=\sigma_{2}(D(S))$ for some $\sigma_{2} \in \mathscr{D}_{n}$.

Second, it is convenient to associate with $S=\left(s_{1}, s_{2}, \ldots, s_{n}\right)$ an $n$-tuple $\mathscr{M}(S) \in \mathscr{S}(1)$ which is related to the parity of the entries of $S$. We define $\mathscr{M}(S)=\left(m_{1}, m_{2}, \ldots, m_{n}\right)$ in the obvious way with $m_{i} \equiv s_{i}(\bmod 2)$. Observe that
(5)

$$
\mathscr{M}(D(S))=D(\mathscr{M}(S))
$$

Theorem 3: Let $n=k r$, where $k=2^{k}$ and $r$ is odd with $r>1$. Suppose for $m \geq 4, \mathscr{L}_{n}(m)=(m-1)(n-2)+k$. Then, for some $\sigma \in \mathscr{D}_{n}$, $D^{2(n-2)}(\mathbb{E})=\sigma(\mathbb{E})$, where $\mathbb{E}=(0, \ldots, 0,1,1)$.
Proof: By hypothesis, there exists an $n$-tuple $S$ with $|S|=m$ and

$$
L(S)=(m-1)(n-2)+k
$$

Let $T=D^{(m-4)(n-2)+1}(S)$. Corollary 1 implies that
$|T|=\left|D^{(m-4)(n-2)+1}(S)\right|=4, \mu(T)=n-2$, and $\left|D^{(n-2)}(T)\right|=3$.
Since $\mu(T)=n-2, T$ has exactly two adjacent entries with values in $\{1,2,3\}$. One of these must equal either 1 or 3 ; for, if not, then $\left|D^{(n-2)}(T)\right| \leq 2$. Moreover, since $T$ has a predecessor, Theorem $1(i)$ guarantees that both are in $\{1,3\}$. This shows that

$$
\mathscr{M}(T)=\sigma_{1}(\mathbb{E}) \text { for some } \sigma_{1} \in \mathscr{D}_{n} .
$$

Similarly,

$$
\mathscr{M}\left(D^{2(n-2)}(T)\right)=\sigma_{2}(\mathbb{E}) \text { for } \sigma_{2} \in \mathscr{D}_{n} .
$$

Hence,

$$
\begin{aligned}
\sigma_{2}(\mathbb{E}) & =\mathscr{M}\left(D^{2(n-2)}(T)\right) \\
& =D^{2(n-2)}(\mathscr{M}(T)) \\
& =D^{2(n-2)}\left(\sigma_{1}(\mathbb{E})\right) \\
& =\sigma_{3}\left(D^{2(n-2)}(\mathbb{E})\right)
\end{aligned}
$$

The second equality follows from (5); the last, from (4). Thus, for $\sigma=\sigma_{3}^{-1} \sigma_{2} \in$ $\mathscr{D}_{n}, D^{2(n-2)}(\mathbb{E})=\sigma(\mathbb{E}) . \quad \square$

Theorem 3 is the heart of the matter. Whether or not there exists an $n-$ tuple which has the maximum possible length depends in large part on $\mathbb{E}$. Since $\mathbb{E} \in \mathscr{S}(1)$, Theorem 3 can be recast in terms of polynomials in $\mathbb{A}$. Using (3), we see that, in order to have an $n$-tuple of maximum length,

$$
(1+t)^{2(n-2)} \mathscr{P}_{\mathbb{E}}(t)=\mathscr{P}_{\sigma(\mathbb{E})}(t)
$$

Since $\mathscr{P}_{\mathbb{E}}(t)=1+t, \mathscr{P}_{\sigma(\mathbb{E})}(t)=t^{j}+t^{j+1}$ for some $j$, where, if necessary, the exponent $j+1$ is reduced modulo $n$. Thus, we have
Corollary 2: Let $n=k r$, where $k=2^{k}$ and $r$ is odd with $r>1$. Suppose that, for $m \geq 4, \mathscr{L}_{n}(m)=(m-1)(n-2)+k$. Then, for some $j$,
(6) $\quad(1+t)^{2 n-3}=t^{j}(1+t)$
in $\mathbb{A}$.
Theorem 4: Let $n$ be an integer such that $n \neq 2^{w}$ and $n \neq 2^{w}+1$ for any $w$. Then, for $m \geq 4, \mathscr{L}_{n}(m)<(m-1)(n-2)+k$.
Proof: First, suppose $n$ is even. By Theorem 1 (ii), $\mathbb{E}=(0, \ldots, 0,1,1)$ is not in a cycle. Thus, $D^{i}(\mathbb{E}) \neq \mathbb{E}$ for any $i$. Now, if $D^{2(n-2)}(\mathbb{E})=\sigma(\mathbb{E})$ for some $\sigma \in \mathscr{D}_{n}$, then $D^{2(n-2) p}(\mathbb{E})=\mathbb{E}$ where $p$ is the order of $\sigma$ in $\mathscr{D}_{n}$. Consequently, the conclusion of Theorem 3 cannot hold.

For $n$ odd, we will expand $(1+t)^{2 n-3}$ denoting the $l^{\text {th }}$ binomial coefficient by $c_{\ell}$.

$$
\begin{aligned}
(1+t)^{2 n-3} & =\sum_{\ell=0}^{2 n-3} c_{\ell} t^{\ell}=\sum_{\ell=0}^{n-3}\left(c_{\ell}+c_{\ell+n}\right) t^{\ell}+\left(c_{n-2} t^{n-2}+c_{n-1} t^{n-1}\right) \\
& =\sum_{l=0}^{n-3}\left(c_{\ell}+c_{n-3-\ell}\right) t^{\ell}+\left(c_{n-2} t^{n-2}+c_{n-2} t^{n-1}\right) \\
& =\sum_{l=0}^{\frac{n-5}{2}}\left(c_{\ell}+c_{n-3-\ell}\right)\left(t^{\ell}+t^{n-3-\ell}\right) \\
& +2 c_{\frac{n-3}{}} t^{\frac{n-3}{2}}+c_{n-2}\left(t^{n-2}+t^{n-1}\right)
\end{aligned}
$$

The second equality follows by using $t^{n}=1$; the third, from $c_{\ell}=c_{2 n-3-\ell}$. Now when $2 n-3=2^{v}-1$ for some $v$, all the binomial coefficients are odd, so that we have

$$
(1+t)^{2 n-3}=t^{n-2}(1+t)
$$

Thus, (6) holds for $n=2^{w}+1$, where $w=v-1$. On the other hand, when $2 n-3 \neq 2^{v}-1$ for any $v$, then $c_{n-2}$ is even. So, if $t^{\ell}$ is present in the expansion of $(1+t)^{2 n-3}$, then so also is $t^{n-3-l}$. Hence, (6) cannot hold.

## 3. The Case $n=2^{w}+1$

We now consider the case in which $n=2^{w}+1$. Corollary 2 and Theorem 4 imply that a game of maximum length is possible. We show that this actually occurs. Before examining the general case, we consider the special case $n=3$.
Lemma 4: Let $n=3$ and define $T_{m}=(m-1,1, m)$. Then, for $m \geq 2, D\left(T_{m}\right)=$ $\sigma\left(T_{m-1}\right)$ for some $\sigma \in \mathscr{D}_{3}$.
Proof: The result is immediate since $D\left(T_{m}\right)=(m-2, m-1,1) . \square$
Lemma 5: Suppose $n=2^{w}+1, w \geq 2$. Let $T_{m}=(0,0, \ldots, 0, m-1,1, m)$. Then, for $m \geq 2$,

$$
\begin{aligned}
& D^{n-4}\left(T_{m}\right)=\left(0, m-1, t_{3}, t_{4}, \ldots, t_{n-1}, m\right) \\
& D^{n-3}\left(T_{m}\right)=(m-1,1, \ldots, 1, m) \\
& D^{n-2}\left(T_{m}\right)=(m-2,0, \ldots, 0, m-1,1)
\end{aligned}
$$

where the entries in $D^{n-4}\left(T_{m}\right)$ have the property that $\left|t_{i}-t_{i+1}\right|=1$ for $i=2$, ..., $n$ - 1 .

Proof: The proof proceeds by induction on $w$. Suppose that $w=2$ so that $n=5$. Then $T_{m}=(0,0, m-1,1, m)$ and it is easily seen that $D\left(T_{m}\right)=(0, m-1$, $m-2, m-1, m)$.

Suppose that the Lemma 5 holds for $w-1$; more specifically, suppose that

$$
\begin{aligned}
& D^{\ell-4}\left(\bar{T}_{m}\right)=\left(0, m-1, r_{3}, r_{4}, r_{\ell-1}, m\right), \text { and } \\
& D^{\ell-4}\left(\bar{T}_{m-1}\right)=\left(0, m-2, s_{3}, s_{4}, \ldots, s_{\ell-1}, m-1\right),
\end{aligned}
$$

where $\ell=2^{\omega-1}+1,\left|r_{i}-r_{i+1}\right|=1$, and $\left|s_{i}-s_{i+1}\right|=1$ for $i=2, \ldots, \ell-1$. Consider the $\left(2^{\omega}+1\right)$-tuple $T_{m}$. We can view $T_{m}$ as a $2^{\omega-1}$ zero-tuple concatenated with a $2^{\omega-1}+1$ " $T_{m}$-type-tuple." Thus, when we compute $D^{k}\left(\mathbb{T}_{m}\right)$ for $k$ less than $2^{\omega-1}-2$, we have the same pattern we have for the $2^{\omega-1}+1$ case. Thus, we have

| $D(T)$ <br> $2^{w-1}-3$ | $\left(0,0, \ldots, 0,0,0, m-1, r_{3}, r_{4}, \ldots, r_{l-1}, m\right)$ |
| :---: | :---: |
| $2^{w-1}-2$ | $(0,0, \ldots, 0,0, m-1,1,1,1, \ldots, 1, m)$ |
| $2^{w-1}-1$ | $(0,0, \ldots, 0, m-1, m-2,0,0, \ldots, 0, m-1, m)$ |
| $2^{w-1}$ | $(0,0, \ldots, m-1,1, m-2,0, \ldots, 0, m-1,1, m)$ |
| $\vdots$ | $\left(0, m-1, s_{l-1}, \ldots, s_{3}, m-2, m-1, r_{3}, r_{4}, \ldots\right.$, |
| $2^{w-3}$ |  |
|  |  |

Note that for $k=2^{w-1}, D^{k}\left(T_{m}\right)$ may be viewed as the $2^{w-1}+1$ " $T_{m-1}$-type" tuple, $(0, \ldots, 0, m-1,1, m-2)$, concatenated with the $2^{w-1}$ tuple, $(0, \ldots, 0, m-1$, $1, m)$. The latter is like the $2^{w-1}+1$ " $T_{m}$-type" tuple, except that it is missing the leading zero. By induction, the second through ( $n-1)^{\text {st }}$ entries in $D^{k}\left(T_{m}\right), k=2^{w}-3=n-4$, differ from the next one by 1. Thus, $D^{n-4}\left(T_{m}\right)$ has the proper form. The conclusion for $D^{n-3}\left(T_{m}\right)$ and $D^{n-2}\left(T_{m}\right)$ follows immediately.
Theorem 5: Suppose $n=2^{w}+1$ for $w \geq 1$. Define $R_{m}=(0,0, \ldots, 0, m-1, m)$ for $m \geq 1$. Then $L\left(R_{m}\right)=(m-1)(n-2)+1$.
Proof: Note that $D\left(R_{m}\right)=T_{m}=(0, \ldots, 0, m-1,1, m)$. Now, by Lemmas 4 and 5, $D^{n-2}\left(T_{m}\right)=\sigma\left(T_{m-1}\right)$ for some $\sigma \in \mathscr{D}_{n}$ and $m \geq 2$. Further, $T_{1}$ is contained in a cycle, but no other $T_{m}$ is. Thus, we have $L\left(R_{m}\right)=(m-1)(n-2)+1$.

## 4. Remaining Questions

For $n$ not a power of 2 and $n \neq 2^{w}+1$, how large is $\mathscr{L}_{n}(m)$ ? What tuple produces the longest game? Only for $n=7$ are the answers to these questions known [6].

Because Theorem 3 cannot hold for even $n$, it is tempting to try to prove a related version using $\mathbb{E}=(0, \ldots, 0,1,0, \ldots, 0,1)$, where the 1 's occur in the $(n-k)^{\text {th }}$ and $n^{\text {th }}$ places. All efforts to date have failed. What relation, if any, does $\mathscr{L}_{2 n}(m)$ have to $\mathscr{L}_{n}(m)$ ? The following is a limited answer to that question.
Theorem 6: $2 \mathscr{L}_{n}(m) \leq \mathscr{L}_{2 n}(m)$.
Proof: Let $S \in \mathscr{S}_{n}(m)$ with $L(S)=\mathscr{L}_{n}(m)$. Then the tuple $S \wedge 0$, where

$$
S \wedge 0=\left(0, s_{1}, 0, s_{2}, 0, s_{3}, \ldots, 0, s_{n}\right)
$$

is in $\mathscr{S}_{2 n}(m)$. By Theorem $1(\mathrm{ii}), D(S \wedge 0)$ is in a cycle if and only if $S$ is. Further, $D^{2}(S \wedge 0)=D(S) \wedge 0$. Thus, $L(S \wedge 0)=2 L(S)$.

Unfortunately, from the few cases studied, it appears that the above inequality is a strict one.

The $n$-number game has been studied extensively; indeed, many key results keep reappearing in the literature and being reproved. An extensive bibliography appears in [7]. In the interest of completeness, additional references which either do not appear in that article or were published after 1982 are listed below.

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TILING THE $k^{\text {th }}$ POWER OF A POWER SERIES<br>Joseph Arkin<br>197 Old Nyack Turnpike, Spring Valley, NY 10977<br>David C. Arney<br>United States Military Academy, West Point, NY 10996<br>Gerald E. Bergum<br>South Dakota State University, Brookings, SD 57007<br>Stefan A. Burr<br>City College, City University of New York, NY 10031<br>Bruce J. Porter<br>United States Military Academy, West Point, NY 10996<br>(Submitted August 1988)

In this paper we consider the problem of dividing a rectangle into nonoverlapping squares and rectangles using recurring-sequence tiling. The results obtained herein are illustrated with appropriate figures. These results, with the exception of basic introductory material, are believed to be new. There seem to be no such results in the literature.

Among the many generating functions possible, we use the following:

$$
\begin{align*}
& G(x)^{-k}=1 /\left(1-x-x^{2}-\ldots-x\right)  \tag{1}\\
& \text { (where } m=2,3,4, \ldots, \text { and } k=1,2,3, \ldots) .
\end{align*}
$$

Note that we can write $G(x)^{-k}$ as a power series in $x$ in the form

$$
\begin{align*}
& G(x)^{-k}=F_{m, k}(0)+F_{m, k}(1) x+F_{m, k}(2) x^{2}+\cdots+F_{m, k}(n) x^{n}+\cdots  \tag{2}\\
& \left(\text { where } F_{m, k}(0)=1, \text { for all } m \text { and } k\right) \text {. }
\end{align*}
$$

We develop a general construction method for performing the tiling using $k \geq 2$. Our work is an extension of the tiling done in [7] for $k=1$.

Using (2),

$$
\begin{align*}
G(x)^{-(k-1)}=\left(1-x-x^{2}\right. & \left.-\cdots-x^{m}\right)\left(F_{m, k}(0)+F_{m, k}(1) x+F_{m, k}(2) x^{2}\right.  \tag{3}\\
& \left.+\cdots+F_{m, k}(n) x^{n}+\cdots\right) .
\end{align*}
$$

Now, combining coefficients in Equation (3) leads to

$$
\begin{equation*}
F_{m, k}(n)=F_{m, k}(n-1)+F_{m, k}(n-2)+\cdots+F_{m, k}(n-m)+F_{m, k-1}(n) . \tag{4}
\end{equation*}
$$

The last term of Equation (4) is important. To preserve the geometry of the method of tiling we have used in this paper, it is necessary that

$$
\begin{equation*}
F_{m, k-1}(n)<F_{m, k}(n-m), \tag{5}
\end{equation*}
$$

where $F_{m, k-1}(n)$ is the value of the initial tile placed in the construction.
First let $m=2$; we shall find the sizes of tiles corresponding to values of $F_{2, k}(n)$ for various values of $k$. In Table 1 , we have outlined the value of the initial tiles generated by the necessary condition that

$$
F_{m, k-1}(n)<F_{m, k}(n-m)
$$

as discussed above. For example, note that $420<474,23109<25088$, etc.
The values in Table 1 can be used in an example tiling construction for $m=2$ and $k=2$, as shown in Figure 2. Note that each shape is a square.

For higher-order constructions, the difficulty lies in choosing the initial tile. We now concern ourselves with finding the value of that required initial tile. A repeated use of Equation (4) above, in example cases of $m=2$ and $m=$ 3 , is summarized in Tables 3 and 4 .

TABLE 1. Coefficients $F_{k}(n)$ when $m=2$

| $n$ | $k=1$ | $k=2$ | $k=3$ | $k=4$ | $k=5$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1 | 1 | 1 | 1 | 1 |
| 1 | 1 | 2 | 3 | 4 | 5 |
| 2 | 2 | 5 | 9 | 14 | 20 |
| 3 | 3 | 10 | 22 | 40 | 65 |
| 4 | 5 | 20 | 51 | 105 | 190 |
| 5 | 8 | 38 | 111 | 256 | 519 |
| 6 | 13 | 71 | 233 | 594 | 1295 |
| 7 | 21 | 130 | 474 | 1324 | 3130 |
| 8 | 34 | 235 | 942 | 2860 | 7285 |
| 9 | 55 | 420 | 1836 | 6020 | 16435 |
| 10 | 89 | 744 | 3522 | 12402 | 36122 |
| 11 | 144 | 1308 | 6666 | 25088 | 77645 |
| 12 | 233 | 2285 | 12473 | 49963 | 163730 |
| 13 | 377 | 3970 | 23109 | 98160 | 339535 |
| 14 | 610 | 6865 | 42447 | 190570 | 693835 |
| 15 | 987 | 11822 | 77378 | 366108 | 1399478 |
| 16 | 1597 | 20284 | 140109 | 696787 | 2790100 |
| 17 | 2584 | 34690 | 252177 | 1315072 | 5504650 |
| 18 | 4181 | 59155 | 451441 | 2463300 | 10758050 |
| 19 | 6765 | 100610 | 804228 | 4582600 | 20845300 |
| 20 | 1096 | 17071 | 1426380 | 847280 | 4007530 |



FIGURE 2

TABLE 3. Calculating the Initial Tile for $m=2$

| $k=2$ | $F_{2,1}(5)$ | $<F_{2,2}(3)$ |
| :--- | :--- | :--- |
| $k=3$ | $F_{2,2}(9)$ | $<F_{2,3}(7)$ |
| $k=4$ | $F_{2,3}{ }^{(13)}$ | $<F_{2,4}(11)$ |
| $k=5$ | $F_{2,4}(17)$ | $<F_{2,5}(15)$ |
| $k=6$ | $F_{2,5}(21)$ | $<F_{2,6}(19)$ |

TABLE 4. Calculating the Initial
Tile for $m=3$

Surveying the values above suggests the following relationships. Define $Q_{m, k}$ to be the smallest number such that

$$
F_{m, k-1}\left(Q_{m, k}+m\right)<F_{m, k}\left(Q_{m, k}\right)
$$

and let $Q_{m}=Q_{m, 1}$. Let $Q_{0}=Q_{1}=1$. Then from Table 3 , where $m=2$, we observe that
(6)

$$
Q_{2}=3 Q_{1}-2 Q_{0}+2 .
$$

Further, we observe from Table 4, where $m=3$, that

$$
\begin{equation*}
Q_{3}=3 Q_{2}-2 Q_{1}+2 . \tag{7}
\end{equation*}
$$

We then generalize that

$$
\begin{equation*}
Q_{m}=3 Q_{m-1}-2 Q_{m-2}+2 \tag{8}
\end{equation*}
$$

By elementary means we find that Equation (8) can be stated as

$$
\begin{equation*}
Q_{m}=2^{m+1}-2 m-1 \tag{9}
\end{equation*}
$$

Then $F_{m, k}\left(Z_{m}\right)$ is the initial tile $I$, where $I=Z_{m}=2^{m+1}-m-1$, for $m \geq 2$.
An examination of Tables 3 and 4 will show that a pattern emerges as $k$ changes and one looks for a value of $n$ which will result in the next initial tile value. As $k$ changes by one, the value of $n$ changes by a constant amount. That constant is equal to
(10) $\quad P_{m}=2^{m+1}-m-2$.

It can be shown inductively, step by step, that the values of the initial tiles are

$$
\begin{equation*}
F_{m, k-1}\left[Z_{m}+(k-2) P_{m}\right]<F_{m, k}\left[Q_{m}+(k-2) P_{m}\right], \text { where } m \geq 2 \text { and } k \geq 1 \tag{11}
\end{equation*}
$$

Then, for example, the next tile values in Table 3, using (11), are seen to be

$$
F_{2,6}(25)<F_{2,7}(23) .
$$

Finally, we now show the general case of placing tiles in the construction. We place the sets of squares in the order Set 1 , Set 2 , etc., where the Sets are defined as


Using this method of set placement, the final general construction will look like Figure 5 below. Note that all shapes are squares, and that as of yet we have not fully tiled the rectangle; there are still rectangular gaps in the construction.

We now proceed to find the filler rectangles used to fill in the gaps left after placing the square tiles. We note first that the general coefficients $F_{m, k}(n)$ may be listed as follows, where $m \geq 2, k \geq 1$, and $n \geq 0$.

$$
\begin{align*}
& F_{m, k}(0)  \tag{16}\\
& F_{m, k}(1) \\
& F_{m, k}(2) \\
& \ldots \\
& F_{m, k}(I)
\end{align*}
$$

where $F_{m, k}(I)=F_{m, k}(I-1)+F_{m, k}(I-2)+\cdots+F_{m, k}(I-m+1)+F_{m, k}(I-m)$ $+F_{m, k-1}(I) ; I=Z_{m}+(k-1) P_{m} ; F_{m, 0}(I)=0, F_{m, k}(0)=1$, and $F_{m, k}(1)=k$.

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FIGURE 5
Note that $I$ is the initial tile value, and it is evident that the value of $I$ depends on the values of $k$ and $m$.

First we examine the case of $m=2$ and $k=2$ [in other words, $1 /(1-x-$ $\left.x^{2}\right)^{2}$ ]. Figure 2 may now be redrawn using function notation rather than actual numbers, and showing the gap rectangles. We use the notation $H$ and $V$ to denote filler rectangles that appear to be oriented horizontally and vertically, respectively.


FIGURE 6

One can see the following rectangle sizes from Figure 6.
(17)

$$
\begin{aligned}
& H_{1, k}\left[F_{2,2}(5)-F_{2,1}(6), F_{2,1}(6)\right] \\
& H_{2, k}\left[F_{2,2}(7)-F_{2,1}(8), F_{2,1}(8)\right] \\
& \cdots \\
& H_{n, k}\left[F_{2,2}(2 n+5)-F_{2,1}(2 n+6), F_{2,1}(2 n+6)\right] \\
& V_{1, k}\left[F_{2,1}(5), F_{2,2}(4)-F_{2,1}(5)\right] \\
& V_{2, k}\left[F_{2,1}(7), F_{2,2}(6)-F_{2,1}(7)\right] \\
& \cdots \\
& V_{n, k}\left[F_{2,1}(2 n+3), F_{2,2}(2 n+2)-F_{2,1}(2 n+5)\right], \\
& \text { where } n \geq 0, m=2, \text { and } k=2
\end{aligned}
$$

We can generalize this idea for any $k$. To use the $I$ notation, Figure 6 may be redrawn as follows:


FIGURE 7

Furthermore, (17) can be rewritten as

$$
\begin{align*}
& H_{1, k}\left[F_{2, k}(I)-F_{2, k}(I+1), F_{2, k-1}(I+1)\right]  \tag{18}\\
& H_{2, k}\left[F_{2, k}(I+2)-F_{2, k-1}(I+3), F_{2, k-1}(I+3)\right] \\
& \cdots \\
& H_{n, k}\left[F_{2, k}(2 n+I)-F_{2, k-1}(2 n+I+1), F_{2, k-1}(2 n+I+1)\right. \\
& V_{1, k}\left[F_{2, k-1}(I), F_{2, k}(I-1)-F_{2, k-1}(I)\right] \\
& V_{2, k}\left[F_{2, k-1}(I+2), F_{2, k}(I+1)-F_{2, k-1}(I+2)\right] \\
& \cdots \\
& V_{n, k}\left[F_{2, k-1}(2 n+I-2), F_{2, k}(2 n+I-3)-F_{2, k-1}(2 n+I-2),\right.
\end{align*}
$$

$$
\text { where } n \geq 0, m=2, \text { and } k \geq 2
$$

We now show an example for $m>2$, in particular for $m=3, k=2$. In this case, the tiling construction begins with the framework shown in Figure 8, using (4) for recursion and Table 4 to determine the initial tile.


FIGURE 8
The filler rectangles are formed in a similar manner as the case of $m=2$, except an alternating pattern containing two different constructions is formed. The complete construction for $k=2, m=3$ is shown in Figure 9. This same pattern is followed for any $k$, where only the value of the arguments of the function $F$ is changed as influenced by the change in the initial tile.


FIGURE 9
One can see the following rectangle sizes from Figure 9:

$$
\begin{align*}
& H_{1,3}\left[F_{3,2}(12)-F_{3,1}(14), F_{3,1}(14)\right]  \tag{19}\\
& H_{2,2}\left[F_{3,2}(13)+F_{3,2}(14), F_{3,2}(15)-F_{3,2}(14)\right] \\
& \ldots \\
& V_{1,2}\left[F_{3,1}(12), F_{3,2}(11)-F_{3,1}(12)\right] \\
& V_{2,2}\left[F_{3,2}(13)-F_{3,2}(12), F_{3,2}(14)-F_{3,2}(13)\right]
\end{align*}
$$

The pattern of the pair of equations in both the horizontal and vertical rectangles is repeated. These formulas would also be valid for any value of $k$, not just $k=2$.

We can now generalize this idea for any $m$. The pattern is repeated every $m$ - 1 rectangles. Therefore, the difference in the argument where this pattern repeats is $2 m-2$. Figure 10 shows the general construction.


FIGURE 10
The general formulas for the horizontal and vertical rectangles are:

$$
\begin{align*}
& H_{1, k}\left[F_{m, k}(I)-F_{m, k-1}(I+m-1), F_{m, k-1}(I+m-1)\right]  \tag{20}\\
& H_{2, k}\left[F_{m, k}(I+1)+F_{m, k}(I+m-1), F_{m, k}(I+m)-F_{m, k}(I+m-1)\right] \\
& H_{3, k}\left[F_{m, k}(I+2)+F_{m, k}(I+m-1)+F_{m, k}(I+m),\right. \\
& \left.\quad F_{m, k}(I+m+1)-F_{m, k}(I+m)\right] \\
& \ddot{H}_{m-1, k}\left[F_{m, k}(I+m-2)+F_{m, k}(I+m-1)+F_{m, k}(I+m)\right. \\
& \left.\quad+\cdots+F_{m, k}(I+2 m-4), F_{m, k}(I+2 m-3)-F_{m, k}(I+2 m-4)\right]
\end{align*}
$$

1990]

$$
\begin{aligned}
& V_{1, k}\left[F_{m, k}(I), F_{m, k}(I-1)-F_{m, k-1}(I)\right] \\
& V_{2, k}\left[F_{m, k}(I+1)-F_{m, k}(I), F_{m, k}(I+m-1)-F_{m, k}(I+m-2)\right. \\
& \left.\quad-\cdots-F_{m, k}(m+2)-F_{m, k}(m+1)\right] \\
& V_{3, k}\left[F_{m, k}(I+2)-F_{m, k}(I+1), F_{m, k}(I+m)-F_{m, k}(I+m-2)\right. \\
& \left.\quad-F_{m, k}(I+m-3)-\cdots-F_{m, k}(I+2)\right] \\
& \cdots \quad V_{m-2, k}\left[F_{m, k}(I+m-3)-F_{m, k}(I+m-4),\right. \\
& \left.F_{m, k}(I+2 m-5)-F_{m, k}(I+m-2)-F_{m, k}(I+m-3)\right] \\
& V_{m-1, k}\left[F_{m, k}(I+m-2)-F_{m, k}(I+m-3),\right. \\
& \left.\quad F_{m, k}(I+2 m-4)-F_{m, k}(I+m-2)\right] .
\end{aligned}
$$

Then the pattern of $m$ - 1 rectangles repeats with the argument ( $I$ ) incremented by $2 m-2$.

This construction, shown in Figure 10, generalizes the recurring-sequence tiling for any $k$ and $m$ using the extended Fibonacci numbers generated with (4). The initial tile is selected by the criterion of (5) and the horizontal and vertical filler rectangles have the dimensions described in (20).

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# ON A HOGGATT-BERGUM PAPER WITH TOTIENT FUNCTION APPROACH FOR DIVISIBILITY AND CONGRUENCE RELATIONS 

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(Submitted August 1988)


#### Abstract

During their discussion of divisibility and congruence relations of the Fibonacci and Lucas numbers, Hoggatt \& Bergum found values of $n$ satisfying the congruences $F_{n} \equiv 0(\bmod n)$ or $L_{n} \equiv 0(\bmod n)$. In this connection, Hoggatt \& Bergum's research appears in Theorems 1, 3, 5, 6, and 7 of [4]. The present paper originated on the same lines in search of values of $n$ that satisfy $\phi\left(F_{n}\right)$ $\equiv 0(\bmod n)$ or $\phi\left(L_{n}\right) \equiv 0(\bmod n)$, where $\phi$ is the totient function. Before going into the analysis of the problem, we state some results that will be quoted frequently. (a) (i) For $n>2, \phi(n)$ is even; (ii) if $m \mid n$, then $\phi(m) \mid \phi(n)([3], p p .140-41)$. (b) All odd prime divisors of $L_{2 n+1}$ are of the form $10 m \pm 1$ ([4], p. 193). (c) Every $F_{n}$ with $n>12$ and $L_{n}$ with $n>6$ has at least one primitive prime divisor ([6], p. 15). (d) Let $p$ be a primitive prime divisor of $F_{n}(n>5)$; if $n \equiv 5(\bmod 10)$, then $p \equiv 1(\bmod 4 n)([6], p .10)$. (e) A primitive prime divisor $p$ of $L_{5 n}$ with $n \geq 1$ satisfies $p \equiv 1$ (mod $10 n$ ) ([6], p. 11). (f) Let $n$ be odd and $p$ an odd primitive prime divisor of $F_{n}$; if $p \equiv \pm 1$ (mod 10), then $p \equiv 1(\bmod 4 n)([1], p .254)$. (g) Let $p$ be an odd primitive prime divisor of $L_{n}$; if $p \equiv \pm 1(\bmod 10)$, then $p \equiv 1(\bmod 2 n)([1], p .255)$.

We begin our discussion by proving the following theorem.


Theorem 1: If $n$ is an odd integer greater than 3 , then
(i) $\phi\left(L_{n}\right) \equiv 0(\bmod n)$;
(ii) $\phi\left(E_{2 n}\right) \equiv 0(\bmod 2 n)$.

Proof: Both results are true when $n=5$. Thus, we choose $n \geq 7$.
(i) Based on (b) and (c), we have the existence of at least one primitive prime divisor $p$ of $L_{n}$ of the form $10 m \pm 1$. Consequently, by (g):
(1) $\quad p \equiv 1(\bmod 2 n)$.

Since $p \mid L_{n}$ and $\phi(p) \equiv 0(\bmod 2 n)$ is true from (1), we have, using (a),
$\phi\left(L_{n}\right) \equiv 0(\bmod 2 n) \Rightarrow \phi\left(L_{n}\right) \equiv 0(\bmod n)$
(ii) Since $F_{2 n}=F_{n} L_{n}$ and $\phi\left(L_{n}\right) \equiv 0(\bmod 2 n)$, we have $2 n \mid \phi\left(F_{2 n}\right)$.

Note: From the above, with odd $n>3, \phi\left(L_{n}\right) \equiv 0(\bmod 2 n)$ and $\phi\left(F_{2 n}\right) \equiv 0(\bmod$ $4 n$ ) are both true. The second part follows from [5].

Corollary: If $n$ is odd, $n>3$, and $3 \mid n$, then $4 n \mid \phi\left(L_{n}\right)$.
Proof: By Lemma 1 of [4] (p. 193), $4 \mid L_{n}$. From Theorem $1, p \mid L_{n}$, where $p \equiv 1$ (mod $2 n$ ) ; consequently, by (a), $\phi(4 p) \mid \phi\left(L_{n}\right)$. This proves our result.

In regard to Fibonacci numbers with even subscripts, we prove the following theorem.
Theorem 2: The congruence $\phi\left(F_{2 N}\right) \equiv 0(\bmod 2 N)$ is true for all positive integers $N$, except when $N=1,2,3,4,8,16$.

Proof: It is easy to verify that, for $N=1,2,4,8,16$, the congruence $\phi\left(F_{2 N}\right)$ $\equiv N(\bmod 2 N)$ holds and, for $N=3$, the result $\phi\left(F_{6}\right) \equiv 4(\bmod 6)$ is true. Excluding these values, we complete the proof by considering the following cases:

Case 1. If $N$ is odd and greater than 3, the result follows from Theorem 1.
Case 2. For even values of $N$, we discuss the proof in two parts:
Part 1. Let $N=2^{n-1}, n \geq 6$. For $n=6$, the result
$\phi\left(F_{64}\right) \equiv 0(\bmod 64)$
is true (see [1], App. Table). For $n>6$, we apply induction on $n$.
$\phi\left(F_{2^{n}}\right) \equiv 0\left(\bmod 2^{n}\right)$ is true by inductive hypothesis.
$\phi\left(L_{2^{n}}\right) \equiv 0(\bmod 2)$ is true by (a), and
$\phi\left(F_{2^{n+1}}\right)=\phi\left(F_{2^{n}}\right) \phi\left(L_{2^{n}}\right)$; therefore, $\phi\left(F_{2^{n+1}}\right) \equiv 0\left(\bmod 2^{n+1}\right)$.
Part 2. Let $N=2^{n-1} t$, where $n \geq 1$, $t$ is odd, $t>1$, and $2 N \neq 6$. If $n=1$, see Theorem 1 . If $n>1$, we use induction on $n$. $F_{2^{n+1} \cdot t}=F_{2^{n} \cdot t L_{2^{n} \cdot t}} ; L_{2^{n}} \mid L_{2^{n} \cdot t}$ and $L_{2^{n}} \& F_{2^{n} \cdot t}$ are relatively prime; therefore,
$\phi\left(F_{2^{n} \cdot t}\right) \cdot \phi\left(L_{2^{n}}\right)$ divides $\phi\left(F_{2^{n+1} \cdot t}\right)$.
Repeating the argument of Part 1 above, we observe that $\phi\left(F_{2^{n} \cdot t}\right) \equiv 0\left(\bmod 2^{n} \cdot t\right)$ is true by the inductive hypothesis, $\phi\left(L_{2^{n}}\right) \equiv 0(\bmod 2)$
follows from (a). Hence, the proof is complete.
For examination of Lucas numbers with even subscripts, it is important to study the values of $\phi\left(L_{2^{n}}\right)$. By verification, it follows that $\phi\left(L_{2^{n}}\right) \equiv 0$ (mod $2^{n}$ ) is true when $n=1,5,6,7,8$ and false for $n=2,3,4$. It remains an open question whether $\phi\left(L_{2^{n}}\right) \equiv 0\left(\bmod 2^{n}\right)$ would be true for all $n \geq 5$ or for infinitely many $n$ or only for a finite number $n$.

Since $2^{n} \equiv 0(\bmod 4)$ is true for $n \geq 5$, every odd prime divisor of $L_{2^{n}}$ is one of the forms $40 m+1,40 m+7,40 m+9,40 m+23$ ([6], p. 11). In [7] it is proved that $L_{2^{n}} \equiv 3$ (mod 4) and, hence, contains an odd number of primes of the form $4 m+3$.

In view of this, we conclude that, if $L_{2} n$ is the product of an even number of primes, then it must contain at least one prime $p$ of the type $40 \mathrm{~m}+1$ or 40 m + 9. If this prime $p$ is primitive, then $p \equiv 1\left(\bmod 2^{n+1}\right)$ by (g). In this case, $2^{n} \mid \phi(p)$ and, consequently, $\phi\left(L_{2^{n}}\right) \equiv 0\left(\bmod 2^{n}\right)$ is true.

Based on this discussion, we are led to make the following conjecture.
Conjecture: There may exist infinitely many $n$ such that $\phi\left(L_{2^{n}}\right) \equiv 0\left(\bmod 2^{n}\right)$ is true.

It is interesting to note that the following allied result holds.
Theorem 3: For all positive integers $n, \phi\left(L_{2^{n}}+1\right) \equiv 0\left(\bmod 2^{n}\right)$ is true.
Proof: Using the Binet form, it is easy to see that

$$
L_{2^{n+1}}+1=\left(L_{2^{n}}+1\right)\left(L_{2^{n}}-1\right)
$$

Since $\left(L_{2^{n}}-1\right)$ is always even, it follows by induction that $L_{2^{n}}+1 \equiv 0$ (mod $\left.2^{n+1}\right)$ is always true. Hence, $2^{n}=\phi\left(2^{n+1}\right)$ divides $\phi\left(L_{2^{n}}+1\right)$.

Some special cases are discussed in the following theorems.
Theorem 4: If $n$ is a positive integer, the following congruences hold:
(i) $\phi\left(L_{5 n}\right) \equiv 0(\bmod 10 n)$.
(ii) $\phi\left(F_{5 n}\right) \equiv 0(\bmod 80 n) ; n$ is odd, $n>1$.

Proof:
(i) The proof follows from (c) and (e) and the fact that $\phi\left(L_{5}\right) \equiv 0$ (mod 10) when $n=1$.
(ii) Since, for odd $n, 5 n \equiv 5(\bmod 10)$ is true when $n>1$, by (c) and (d) there exists a primitive prime divisor $p$ of $F_{5 n}$ satisfying $p \equiv 1$ (mod 20n). Since 5 and $p$ are both relatively prime factors of $F_{5 n}, 80 n$ divides $\phi\left(F_{5 n}\right)$ by (a).
Theorem 5:
(i) If $k \geq 2$, then $\phi\left(F_{5^{k}}\right) \equiv 0\left(\bmod 16 \cdot 5^{2 k-1}\right)$.
(ii) If $n=2^{r+1} \cdot 3^{m} \cdot 5^{k}$ with $r \geq 1, m \geq 1, k \geq 1$, then $\phi\left(F_{n}\right) \equiv 0$ (mod $\left.4 n\right)$. Proof:
(i) From [4], p. 192, we have $5^{k} \mid F_{5^{k}}$. Since $5^{k} \equiv 5(\bmod 10)$ is true, by (d) there exists a primitive prime divisor $p$ satisfying $p \equiv 1$ (mod $\left.4.5^{k}\right)$. As $5^{k}$ and $p$ are relatively prime, $\phi\left(5^{k}\right) \cdot \phi(p)$ divides $\phi\left(F_{k}\right)$.
(ii) By [4], p. 192, we have $n \mid F_{n}$. This, along with Theorem 2 , completes the proof.
A Final Note: It is desirable to shed some light on the cases not discussed thus far and on the difficulties encountered in the generalization process. This is done by showing that the following two congruences are not valid in general for a positive integer $n$.
(i) $\phi\left(L_{2 n}\right) \equiv 0(\bmod 2 n)$
(ii) $\phi\left(F_{2 n+1}\right) \equiv 0(\bmod 2 n+1)$

In regard to (i), we observe that if, for a composite $m$, $L_{m}$ is prime, then $m$ must be of the form $2^{t}$, where $t \geq 2$ (see [2]). Consequently, with $t \geq 2$ when $L_{2}^{t}$ is prime, which is primitive, we have $\phi\left(L_{2^{t}}\right)=L_{2}^{t}-1$.

As proved in Theorem $3, L_{2} t \equiv-1\left(\bmod 2^{t}\right)$; therefore, we can conclude that $\phi\left(L_{2}\right) \equiv-2\left(\bmod 2^{t}\right)$. Thus, it follows that $\phi\left(L_{2^{t}}\right) \not \equiv 0\left(\bmod 2^{t}\right)$ when $L_{2}$ is prime and $t \geq 2$. Besides this, there may exist other Lucas numbers connected with this which may not satisfy the congruence of (i). One such illustration will be the members of the type $L_{2^{t} . p}$, where $p$ is an odd prime and $t \geq 1$. We observe that, for $n<50, \phi\left(L_{2 n}\right) \not \equiv 0(\bmod 2 n)$ when $n=2,4,8,11,12,17,26$, 29, 37, 46. In view of this, we conclude that the congruence relation in (i) is not true in general.

For case (ii), we observe that for odd $n$, if $F_{n}$ is prime $p$, where $p \equiv \pm 3$ $(\bmod 10)$, then $p \equiv(2 n-1)(\bmod 4 n)($ see $[1], p .254)$.

Thus, under this hypothesis of primality, $\phi\left(F_{n}\right) \equiv-2(\bmod n)$. It is easy to see that $F_{n}$ is a prime of this type when $n=7,13,17,23,43,47,83$. It is interesting to observe that if, for a prime subscript $n, F_{n}$ is the product of two primitive primes each $\equiv \pm 3(\bmod 10)$, then $\phi\left(F_{n}\right) \equiv 4(\bmod n)$. This is true when $n=59,61,71,79,101,109$.

Based on this, there may exist Fibonacci numbers of odd subscripts $n$, where $n$ is composite, which may also not satisfy relation (ii). One such example is $F_{161}$, where $\phi\left(F_{161}\right) \equiv 16$ (mod 161). Consequently, we are justified to say that the congruence relation of (ii) is also not true in general.

## References

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## ELEMENTARY PROBLEMS AND SOLUTIONS

Edited by
A. P. Hillman

Please send all material for ELEMENTARY PROBLEMS AND SOLUTIONS to Dr. A. P. HILLMAN; 709 SOLANO DR., S.E.; ALBUQUERQUE, NM 87108.

Each solution should be on a separate sheet (or sheets) and must be received within six months of publication of the problem. Solutions typed in the format used below will be given preference. Proposers of problems should include solutions.

Anyone desiring acknowledgment of contributions should enclose a stamped, self-addressed card (or envelope).

## BASIC FORMULAS

The Fibonacci numbers $F_{n}$ and the Lucas numbers $L_{n}$, satisfy
$F_{n+2}=F_{n+1}+F_{n}, F_{0}=0, F_{1}=1 ;$
$L_{n+2}=L_{n+1}+L_{n}, L_{0}=2, L_{1}=1$.
Also, $\alpha=(1+\sqrt{5}) / 2, \beta=(1-\sqrt{5}) / 2, F_{n}=\left(\alpha^{n}-\beta^{n}\right) / \sqrt{5}$, and $L_{n}=\alpha^{n}+\beta^{n}$.

## PROBLEMS PROPOSED IN THIS ISSUE

B-670 Proposed by Russell Euler, Northwest Missouri State U., Marysville, MO Evaluate $\sum_{n=1}^{\infty} \frac{n F_{n}}{2^{n}}$.

B-671 Proposed by Herta T. Freitag, Roanoke, VA
Show that all even perfect numbers are hexagonal and hence are all triangular. [A perfect number is a positive integer which is the sum of its proper positive integral divisors. The hexagonal numbers are $\{1,6,15,28,45, \ldots\}$ and the triangular numbers are $\{1,3,6,10,15, \ldots\}$.

B-672 Proposed by Philip L. Mana, Albuquerque, NM
Let $S$ consist of all positive integers $n$ such that $n=10 p$ and $n+1=11 q$, with $p$ and $q$ primes. What is the largest positive integer $d$ such that every $n$ in $S$ is a term in an arithmetic progression $a, a+d, a+2 d, \ldots$ ?

B-673 Proposed by Paul S. Bruckman, Edmonds, WA
Evaluate the infinite product $\prod_{n=2}^{\infty} \frac{F_{2 n}+1}{F_{2 n}-1}$.
B-674 Proposed by Richard André-Jeannin, Sfax, Tunisia
Define the sequence $\left\{u_{n}\right\}$ by

$$
u_{0}=0, u_{1}=1, u_{n}=g u_{n-1}-u_{n-2}, \text { for } n \text { in }\{2,3, \ldots\},
$$

where $g$ is a root of $x^{2}-x-1=0$. Compute $u_{n}$ for $n$ in $\{2,3,4,5\}$ and then deduce that $(1+\sqrt{5}) / 2=2 \cos (\pi / 5)$ and $(1-\sqrt{5}) / 2=2 \cos (3 \pi / 5)$.

B-675 Proposed by Richard André-Jeannin, Sfax, Tunisia
In a manner analogous to that for the previous problem, show that

$$
\sqrt{2+\sqrt{2}}=2 \cos \frac{\pi}{8} \text { and } \sqrt{2-\sqrt{2}}=2 \cos \frac{3 \pi}{8} .
$$

## SOLUTIONS

## Not True Asymptotically

B-645 Proposed by R. Tošić, U. of Novi Sad, Yugoslavia

$$
\text { Let } \begin{aligned}
G_{2 m} & =\binom{2 m-1}{m}-2\binom{2 m-1}{m-3}+\binom{2 m}{m-5} \text { for } m=1,2,3, \ldots, \\
G_{2 m+1} & =\binom{2 m}{m}-\binom{2 m+1}{m-2}+2\binom{2 m}{m-5} \quad \text { for } m=0,1,2, \ldots,
\end{aligned}
$$

where $\binom{n}{k}=0$ for $k<0$. Prove or disprove that $G_{n}=F_{n}$ for $n=0,1,2, \ldots$.
Solution by Y. H. Harris Kwong, SUNY College at Fredonia, Fredonia, NY
Let us study the asymptotic growth of $G_{n}$. It is evident that

$$
G_{2 m} \sim\binom{2 m-1}{m} \quad \text { and } \quad G_{2 m+1} \sim\binom{2 m}{m} .
$$

Hence,

$$
\frac{G_{2 m+1}}{G_{2 m}} \sim \frac{\binom{2 m}{m}}{\binom{2 m-1}{m}}=\frac{(2 m)!}{m!m!} \cdot \frac{m!(m-1)!}{(2 m-1)!}=2
$$

and

$$
\frac{G_{2 m+2}}{G_{2 m+1}} \sim \frac{\binom{2 m+1}{m+1}}{\binom{2 m}{m}}=\frac{(2 m+1)!}{(m+1)!m!} \cdot \frac{m!m!}{(2 m)!}=\frac{2 m+1}{m+1} \sim 2,
$$

so that $G_{n} / G_{n-1} \sim 2$. However, it is well known that

$$
F_{n} / F_{n-1} \sim(1+\sqrt{5}) / 2
$$

Thus, $G_{n} \neq F_{n}$ for sufficiently large $n$. In fact, from numerical computations, we have $G_{n}=F_{n}$ for $0<n \leq 14$, and $G_{n}>F_{n}$ for $n \geq 15$.

Also solved by Charles Ashbacher, Paul S. Bruckman, James E. Desmond, Piero Filipponi, L. Kuipers, and the proposer.

## Triangular Number Analogue

B-646 Proposed by A. P. Hillman in memory of Gloria C. Padilla
We know that $F_{2 n}=F_{n} L_{n}=F_{n}\left(F_{n-1}+F_{n+1}\right)$. Find $m$ as a function of $n$ so as to have the analogous formula $T_{m}=T_{n}\left(T_{n-1}+T_{n+1}\right)$, where $T_{n}$ is the triangular number $n(n+1) / 2$.

Solution by H.-J. Seiffert, Berlin, Germany

$$
\text { We have: } \begin{aligned}
T_{n}\left(T_{n-1}+T_{n+1}\right) & =T_{n}\left(T_{n}-n+T_{n}+n+1\right)=T_{n}\left(2 T_{n}+1\right) \\
& =n(n+1)(n(n+1)+1) / 2=T_{n(n+1)} .
\end{aligned}
$$

Also solved by Richard André-Jeannin, Wray G. Brady, Paul S. Bruckman, Nicos D. Diamantis, Russell Euler, Piero Filipponi, Herta T. Freitag, Russell Jay Hendel, L. Kuipers, Jack Lee, Carl Libis, Bob Prielipp, Jesse Nemoyer \& Joseph J. Kostal \& Durbha Subramanyam, Sahib Singh, Lawrence Somer, Gregory Wulczyn, and the proposer.

## Much Ado about Zero

B-647 Proposed by L. Kuipers, Serre, Switzerland
Simplify

$$
\left[L_{2 n}+7(-1)^{n}\right]\left[L_{3 n+3}-2(-1)^{n} L_{n}\right]-3(-1)^{n} L_{n-2} L_{n+2}^{2}-L_{n-2} L_{n-1} L_{n+2}^{3} .
$$

Solution by Sahib Singh, Clarion U. of Pennsylvania, Clarion, PA
The given expression simplifies to zero. By using the Binet form of Lucas numbers, it follows that $L_{2 n}+7(-1)^{n}=L_{n-2} L_{n+2}$. In view of this, the given expression is

$$
L_{n-2} L_{n+2}\left[L_{3 n+3}-2(-1)^{n} L_{n}-\left(L_{n+2}^{2} L_{n-1}+3(-1)^{n} L_{n+2}\right)\right]
$$

Again, applying the Binet form of Lucas numbers, we see that

$$
L_{n+2}^{2} L_{n-1}+3(-1)^{n} L_{n+2}=L_{3 n+3}-2(-1)^{n} L_{n} .
$$

Hence, the required conclusion follows.
Also solved by Paul S. Bruckman, Herta T. Freitag, Y. H. Harris Kwong, Carl Libis, Bob Prielipp, H.-J. Seiffert, M. Wachtel, Gregory Wulczyn, and the proposer.

## Pell Primitive Pythagorean Triples

B-648 Proposed by M. Wachtel, Zurich, Switzerland
The Pell numbers $P_{n}$ and $Q_{n}$ are defined by

$$
P_{n+2}=2 P_{n+1}+P_{n}, P_{0}=0, P_{1}=1 ; Q_{n+2}=2 Q_{n+1}+Q_{n}, Q_{0}=1=Q_{1} .
$$

Show that $\left(P_{4 n}, P_{2 n}^{2}+1,3 P_{2 n}^{2}+1\right)$ is a primitive Pythagorean triple for $n$ in \{1, 2, ...\}.

Solution by Paul S. Bruckman, Edmonds, WA
The Pell numbers satisfy the following identities:

$$
\begin{align*}
& 2 P_{2 n} Q_{2 n}=P_{4 n}  \tag{1}\\
& Q_{2 n}^{2}-2 P_{2 n}^{2}=1 \tag{2}
\end{align*}
$$

Hence,

$$
\begin{equation*}
Q_{2 n}^{2}-P_{2 n}^{2}=P_{2 n}^{2}+1 \tag{3}
\end{equation*}
$$

It is known that primitive Pythagoren triples are generated by
(4) (2ab, $\left.a^{2}-b^{2}, a^{2}+b^{2}\right)$, where g.c.d. $(a, b)=1$.

We may let $a=Q_{2 n}, b=P_{2 n}$. We see from (2) that g.c.d. $(\alpha, b)=1$. Also

$$
\begin{aligned}
& 2 a b=P_{4 n} \quad[\text { using (1) }] \\
& a^{2}-b^{2}=P_{2 n}^{2}+1 \quad[\text { using (3) }]
\end{aligned}
$$

and

$$
a^{2}+b^{2}=3 P_{2 n}^{2}+1 \quad\left[\text { adding } 2 b^{2}\right. \text { to both sides of (3)]. }
$$

This proves the assertion.
Also solved by Nicos D. Diamantis, Ernest J. Eckert, Russell Euler, Piero Filipponi, Herta T. Freitag, Russell Jay Hendel, L. Kuipers, Jesse Nemoyer \& Joseph J. Kostal \& Durbha Subramanyam, Bob Prielipp, H.-J. Seiffert, Sahib Singh, Lawrence Somer, Gregory Wulczyn, and the proposer.

## Sides Differing by 17

B-649 Proposed by M. Wachtel, Zurich, Switzerland
Give a rule for constructing a sequence of primitive Pythagorean triples ( $a_{n}, b_{n}, c_{n}$ ) whose first few triples are in the table

| $n$ | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 |
| :--- | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $a_{n}$ | 24 | 28 | 88 | 224 | 572 | 1248 | 3276 | 7332 |
| $b_{n}$ | 7 | 45 | 105 | 207 | 555 | 1265 | 3293 | 7315 |
| $c_{n}$ | 25 | 53 | 137 | 305 | 797 | 1777 | 4645 | 10357 |

and which satisfy

$$
\left|a_{n}-b_{n}\right|=17,
$$

$$
a_{2 n-1}+a_{2 n}=26 P_{2 n}=b_{2 n-1}+b_{2 n},
$$

and

$$
c_{2 n-1}+c_{2 n}=26 Q_{2 n}
$$

[ $P_{n}$ and $Q_{n}$ are the Pell numbers of $B-648$.]
Rule by Paul S. Bruckman, Edmonds, WA

$$
\begin{aligned}
& \left(\alpha_{2 n-1}, b_{2 n-1}, c_{2 n-1}\right) \\
& =\left(10 P_{n}^{2}+26 P_{n} Q_{n}-12 Q_{n}^{2},-24 P_{n}^{2}+26 P_{n} Q_{n}+5 Q_{n}^{2}, 26 P_{n}^{2}-14 P_{n} Q_{n}+13 Q_{n}^{2}\right), \\
& \left(\alpha_{2 n}, b_{2 n}, c_{2 n}\right) \\
& =\left(-10 P_{n}^{2}+26 P_{n} Q_{n}+12 Q_{n}^{2}, 24 P_{n}^{2}+26 P_{n} Q_{n}-5 Q_{n}^{2}, 26 P_{n}^{2}+14 P_{n} Q_{n}+13 Q_{n}^{2}\right) .
\end{aligned}
$$

Rule by Ernest J. Eckert, U. of South Carolina, Aiken, SC
Let $\left(a_{1}, b_{1}, c_{1}\right)=(24,7,25),\left(a_{2}, b_{2}, c_{2}\right)=(28,45,53)$ and $A$ denote the matrix

$$
\left[\begin{array}{lll}
1 & 2 & 2 \\
2 & 1 & 2 \\
2 & 2 & 3
\end{array}\right]
$$

Then $\left[\begin{array}{lll}a_{2 n-1} & b_{2 n-1} & c_{2 n-1}\end{array}\right]$ is the matrix product $\left[\begin{array}{lll}a_{1} & b_{1} & c_{1}\end{array}\right] A^{n-1}$ and

$$
\left[\begin{array}{lll}
a_{2 n} & b_{2 n} & c_{2 n}
\end{array}\right]=\left[\begin{array}{lll}
a_{2} & b_{2} & c_{2}
\end{array}\right] A^{n-1}
$$

Editor's note: The derivations and proofs given by Bruckman and Eckert are not included because of space limitations; however, since each term in the required equations satisfies the same $3^{\text {rd }}$ order linear homogeneous recursion

$$
w_{n+3}=5\left(w_{n+2}+w_{n+1}\right)-w_{n},
$$

it suffices to verify the rules for $n=1,2$, and 3 .
Also solved by Gregory Wulczyn and the proposer.

## Average Age of Generalized Rabbits

B-650 Proposed by Piero Filipponi, Fond. U. Bordoni, Rome Italy \& David Singmaster, Polytechnic of the South Bank, London, UK

Let us introduce a pair of l-month-old rabbits into an enclosure on the first day of a certain month. At the end of one month, rabbits are mature and each pair produces $k-1$ pairs of offspring. Thus, at the beginning of the second month there is 1 pair of 2 -month-old rabbits and $k-1$ pairs of $0-m o n t h-$ olds. At the beginning of the third month, there is 1 pair of 3 -month-olds, $k-1$ pairs of 1 -month-olds, and $k(k-1)$ pairs of 0 -month-olds. Assuming that the rabbits are immortal, what is their average age $A_{n}$ at the end of the $n$th month? Specialize to the first few values of $k$. What happens as $n \rightarrow \infty$ ?

Solution by Sahib Singh, Clarion U. of Pennsylvania, Clarion, PA
If $A_{i}$ denotes the average age at the end of the $i$ th month, then we have the following recurrence relation:

$$
A_{i+1}=\frac{1}{k}\left(1+A_{i}\right), \text { where } A_{1}=\frac{2}{k} ; k>1
$$

Using this, we conclude that

$$
A_{n}=\frac{1}{k^{n}}\left(1+\sum_{i=0}^{n-1} k^{i}\right)=\frac{k^{n}+k-2}{k^{n}(k-1)}
$$

Thus,

$$
\begin{aligned}
& A_{2}=\frac{k+2}{k^{2}} ; \quad A_{3}=\frac{k^{2}+k+2}{k^{3}}, \text { etc. } \\
& \operatorname{Limit}_{n \rightarrow \infty} A_{n}=\frac{1}{k-1} .
\end{aligned}
$$

Also solved by Paul S. Bruckman and the proposers.

## Multiples of a Prime $p$

B-651 Proposed by L. Van Hamme, Vrije Universiteit, Brussels, Belgium
Let $u_{0}, u_{1}, \ldots$ be defined by $u_{0}=0, u_{1}=1$, and $u_{n+2}=u_{n+1}-u_{n}$. Also let $p$ be a prime greater than 3 , and for $n$ in $X=\{1,2, \ldots, p-1\}$, let $n^{-1}$ denote the $v$ in $X$ with $n v \equiv 1(\bmod p)$. Prove that

$$
\sum_{n=1}^{p-1}\left(n^{-1} u_{n+k}\right) \equiv 0(\bmod p)
$$

for all nonnegative integers $k$.
Solution by the proposer.
Let $\rho$ be a zero of $1+X+X^{2}$. Hence, $\rho^{3}=1$. Since

$$
\begin{aligned}
(1+\rho)^{p}-1-\rho^{p} & =-\rho^{2 p}-1-\rho^{p} \\
& =-\left(\rho^{2}+1+\rho\right)=0 \quad \\
& =-\left(\rho^{-2}+1+\rho^{-1}\right)=0
\end{aligned} \quad \begin{aligned}
& \text { if } p \equiv 1(\bmod 3) \\
&
\end{aligned}
$$

$\rho$ is also a zero of $(1+X)^{p}-1-X$. Hence,

$$
\sum_{n=1}^{p-1}\binom{p}{n} \rho^{n}=0, \quad \sum_{n=1}^{p-1}\binom{p}{n} \rho^{-n}=0
$$

Multiplying the first equation with $p k$, the second with $\rho^{-k}$, and using the easily verified formula

$$
u_{n}=\frac{(-1)^{n-1}}{\sqrt{-3}}\left(\rho^{n}-\rho^{-n}\right),
$$

we get

$$
\sum_{n=1}^{p-1}(-1)^{n-1} u_{n+k}\binom{p}{n}=0
$$

Dividing by $p$ and using

$$
\frac{1}{p}\binom{p}{n} \equiv \frac{(-1)^{n-1}}{n} \quad(\bmod p), \quad 1 \leq n \leq p-1
$$

we get the assertion.
Also solved by Paul S. Bruckman.
*****
(continued from page 288)
$Z_{i}(t)$ represents the number of zeros of $f_{t}$ which are $\varepsilon$-close to $\eta_{i}$. By invariance of the complex integral, the functions $Z_{i}(t)$ are constant since the functions $f_{t}$ vary continuously and do not vanish on the path of integration. Hence, $Z_{i}(0)=Z_{i}(1)$ for each $i$. This says that in a small neighborhood of each zero of $f_{1}$, there is a one-to-one correspondence of zeros of $f_{1}$ with zeros of $f_{0}$, in the required manner.

In the case of our given functions, we find that the zeros of the polynomial $f_{n}(z)$ are close to the zeros of $g_{n}(z)$, which lie on the circle $|z|=\alpha$, as required, and the zeros of $f_{n}$ get closer to the circle as $n \rightarrow \infty$.

Also solved by P. Bruckman, O. Brugia \& P. Filipponi, L. Kuipers, and the proposer.

## *****

## ADVANCED PROBLEMS AND SOLUTIONS

Edited by
Raymond E. Whitney
Please send all communications concerning ADVANCED PROBLEMS AND SOLUTIONS to RAYMOND E. WHITNEY, MATHEMATICS DEPARTMENT, LOCK HAVEN UNIVERSITY, LOCK HAVEN, PA 17745. This department especially welcomes problems believed to be new or extending old results. Proposers should submit solutions or other information that will assist the editor. To facilitate their consideration, all solutions should be submitted on separate signed sheets within two months after publication of the problems.

## PROBLEMS PROPOSED IN THIS ISSUE

H-443 Proposed by Richard Andre-Jeannin, Sfax, Tunisia
Let us consider the recurrence

$$
w_{n}=m w_{n-1}+w_{n-2},
$$

where $m>0$ is an integer and $U_{n}, V_{n}$ the solutions defined by

$$
U_{0}=0, U_{1}=1 ; V_{0}=2, V_{1}=m .
$$

Show that, if $q$ is an odd divisor of $m^{2}+1$, then

$$
V_{q} \equiv m(\bmod q) .
$$

H-444 Proposed by H.-J. Seiffert, Berlin, Germany
Show that, for $n=0,1,2, \ldots$,

$$
F_{n}=\sum_{\substack{k=0 \\(5, n-2 k)=1}}^{[n / 2]}(-1)^{[(n-2 k+2) / 5]}\binom{n}{k},
$$

where ( $r, s$ ) denotes the greatest common divisor of $r$ and $s$ and [ ] the greatest integer function.

H-445 Proposed by Paul S. Bruckman, Edmonds, WA
Establish the identity:

$$
\begin{equation*}
\sum_{n=1}^{\infty} \mu(n)\left(\frac{z^{n}}{1-z^{2 n}}\right)=z-z^{2}, z \in \mathbb{C},|z|<1 \text {, and } \mu \text { is the Möbius function. } \tag{1}
\end{equation*}
$$

As special cases of (1), obtain the following identities:

$$
\begin{equation*}
\sum_{n=1}^{\infty} \mu(2 n) / F_{2 n s}=-\beta^{2 s} \sqrt{5}, s=1,3,5, \ldots, \beta=\frac{1}{2}(1-\sqrt{5}) ; \tag{2}
\end{equation*}
$$

3) $\quad \sum_{n=1}^{\infty} \mu(2 n-1) / L_{(2 n-1) s}=-\beta^{s}, s=1,3,5, \ldots$;

$$
\begin{equation*}
\sum_{n=1}^{\infty} \mu(n) / F_{n s}=\left(\beta^{s}-\beta^{2 s}\right) \sqrt{5}, s=2,4,6, \ldots \tag{3}
\end{equation*}
$$

$$
\begin{align*}
& \sum_{n=1}^{\infty}(-1)^{n-1} \mu(n) / F_{n s}=\left(\beta^{s}+\beta^{2 s}\right) \sqrt{5}, s=2,4,6, \ldots ;  \tag{5}\\
& \sum_{n=1}^{\infty}(-1)^{n-1} \mu(2 n-1) / F_{(2 n-1) s}=-\beta^{s} \sqrt{5}, s=1,3,5, \ldots ; \\
& \sum_{n=1}^{\infty}(-1)^{n-1} \mu(2 n-1) / L_{(2 n-1) s}=\beta^{s}, s=2,4,6, \ldots .
\end{align*}
$$

## SOLUTIONS

## Rather Compact

H-421 Proposed by Piero Filipponi, Rome, Italy (Vol. 26, no. 2, May 1988)

Let the numbers $U_{n}(m)$ (or merely $U_{n}$ ) be defined by the recurrence relation [1]

$$
U_{n+2}=m U_{n+1}+U_{n} ; U_{0}=0, U_{1}=1,
$$

where $m \in N=\{1,2, \ldots\}$.
Find a compact form for

$$
S(k, h, n)=\sum_{j=0}^{n-1} U_{k+j h} U_{k+(n-1-j) h}(k, h, n \in N) .
$$

Note that, in the particular case $m=1, S(1,1, n)=F_{n}^{(1)}$ is the $n^{\text {th }}$ term of the Fibonacci first convolution sequence [2].

## References

1. M. Bickne11. "A Primer on the Pell Sequence and Related Sequences." Fibonacei Quarterly 13.4 (1975):345-49.
2. V. E. Hoggatt, Jr. "Convolution Triangles for Generalized Fibonacci Numbers." Fibonacci Quarterly 8.2 (1970):158-71.

Solution by the proposer
It is known [1] that

$$
\begin{equation*}
U_{n}=\left(\alpha^{n}-\beta^{n}\right) / \Delta \quad(\text { Binet form }) \tag{1}
\end{equation*}
$$

where

$$
\left\{\begin{array}{l}
\Delta=\left(m^{2}+4\right)^{1 / 2},  \tag{2}\\
\alpha=(m+\Delta) / 2, \\
\beta=(m-\Delta) / 2 .
\end{array}\right.
$$

Analogously, the numbers $V_{n}(m)$ (or merely $V_{n}$ ) can be defined as either

$$
\begin{equation*}
V_{n+2}=m V_{n+1}+V_{n} ; V_{0}=2, V_{1}=m, \tag{3}
\end{equation*}
$$

or

$$
\begin{equation*}
V_{n}=U_{n-1}+U_{n+1}=\alpha^{n}+\beta^{n} \quad(\text { Binet form }) \tag{4}
\end{equation*}
$$

The following identities will be used throughout the solution:

$$
\begin{align*}
& V_{j+k}-(-1)^{k} V_{j-k}=\Delta^{2} U_{j} U_{k} ;  \tag{5}\\
& \left\{\begin{array}{l}
U_{-n}=(-1)^{n+1} U_{n}, \\
V_{-n}=(-1)^{n} V_{n} ;
\end{array}\right. \tag{6}
\end{align*}
$$

$$
\begin{equation*}
\sum_{j=0}^{r} x^{j} V_{s j+t}=\frac{(-1)^{s-1} x^{r+2} V_{s r+t}+x^{p+1} V_{s(r+1)+t}+(-1)^{t} x V_{s-t}-V_{t}}{(-1)^{s-1} x^{2}+V_{s .} x-1} \tag{7}
\end{equation*}
$$

where $s$ and $t$ are arbitrary integers and $x$ is an arbitrary quantity subject to the restriction $x \neq \alpha^{-s}, \beta^{-s}$.

Identities (5) and (6) can be readily proven with the aid of (1), (2), and (4). The proof of (7) is slightly more complicated but several approaches are possible. One of these proofs is given in "A Matrix Approach to Certain Identities" by P. Filipponi \& A. F. Horadam (Fibonacci Quarterly 26.2 [1988]:11526).

Now, from (5), we can write

$$
U_{k+j h} U_{k}+(n-1-j) h=\left[V_{2 k+(n-1) h}-(-1)^{k+(n-j-1) h} V_{(2 j-n+1) h}\right] / \Delta^{2}
$$

whence

$$
\begin{align*}
S(k, h, n) & =\frac{n V_{2 k+(n-1) h}}{\Delta^{2}}-\frac{(-1)^{k+(n-1) h}}{\Delta^{2}} \sum_{j=0}^{n-1}(-1)^{j h} V_{2 h j-(n-1) h}  \tag{8}\\
& =\frac{n V_{2 k+(n-1) h}}{\Delta^{2}}-\frac{(-1)^{k+(n-1) h}}{\Delta^{2}} X_{h, n}
\end{align*}
$$

Using (7), (6), and (5), let us calculate the quantity $X_{h, n}$ :
Case 1: $h$ is odd [ $x=-1$ in (7)]

$$
\begin{equation*}
X_{n, h}=\frac{2(-1)^{n-1}\left[V_{h(n+1)}+V_{h(n-1)}\right]}{V_{2 h}+2}=\frac{2(-1)^{n-1} \Delta^{2} U_{h n} U_{h}}{V_{2 h}+2} . \tag{9}
\end{equation*}
$$

Using (1) and (4), (9) becomes

$$
\begin{equation*}
x_{h, n}=2(-1)^{n-1} U_{U_{n}} / U_{n} . \tag{10}
\end{equation*}
$$

Case 2: $h$ is even [ $x=1$ in (7)]

$$
\begin{equation*}
X_{h, n}=\frac{2\left[V_{h(n+1)}-V_{h(n-1)}\right]}{V_{2 h}-2}=\frac{2 \Delta^{2} U_{h n} U_{h}}{V_{2 h}-2}=2 U_{h n} / U_{h} . \tag{11}
\end{equation*}
$$

From (8), (9), and (10), we obtain

$$
\begin{equation*}
S(k, h, n)=\left[n V_{2 k}+(n-1) h-2(-1)^{k} U_{h n} / U_{h}\right] / \Delta^{2} \tag{12}
\end{equation*}
$$

The relationship (4) allows us to express $S(k, h, n)$ merely in terms of numbers $U_{n}$ 。

As a particular case, we have
(13) $S(1,1, n)=\left[n V_{n+1}+2 U_{n}\right] / \Delta^{2}=\left[n U_{n+2}+(n+2) U_{n}\right] / \Delta^{2}$.

Also solved by P. Bruckman, L. Kuipers, H.-J. Seiffert, and N. A. Volodin.

## Lotsa Sequences

H-422 Proposed by Larry Taylor, Rego Park, NY (Vol. 26, no. 2, May 1988)
(A1) Generalize the numbers (2, 2, 2, 2, 2, 2, 2) to form a seven-term arithmetic progression of integral multiples of Fibonacci and/or Lucas numbers with common difference $F_{n}$.
(A2) Generalize the numbers (1, 1, 1, 1, 1, 1) to form a six-term arithmetic progression of integral multiples of Fibonacci and/or Lucas numbers with common difference $F_{n}$.
(A3) Generalize the numbers (4, 4, 4, 4, 4) to form a five-term arithmetic progression of integral multiples of Fibonacci and/or Lucas numbers with common difference $5 F_{n}$.
(A4) Generalize the numbers $(3,3,3,3),(3,3,3,3),(3,3,3,3)$ to form three four-term arithmetic progressions of integral multiples of Fibonacci and/or Lucas numbers with common differences $F_{n}, 5 F_{n}, F_{n}$.
(B) Generalize the Fibonacci and Lucas numbers in such a way that, if the Fibonacci numbers are replaced by the generalized Fibonacci numbers and the Lucas numbers are replaced by the generalized Lucas numbers, the arithmetic progressions still hold.

Solution by Paul S. Bruckman, Edmonds, WA
We indicate below the solutions to parts (A1)-(A4) of the problem:

$$
\begin{align*}
& \left(-2 F_{n-2}, F_{n-3}, 2 F_{n-1}, L_{n}, 2 F_{n+1}, F_{n+3}, 2 F_{n+2}\right) \text {; }  \tag{A1}\\
& \left(-L_{n-1},-F_{n-2}, F_{n-1}, F_{n+1}, F_{n+2}, L_{n+1}\right) ;  \tag{A2}\\
& \left(-4 L_{n-1},-L_{n-3}, 2 L_{n}, L_{n+3}, 4 L_{n+1}\right) ;  \tag{A3}\\
& \text { (i) }\left(3 F_{n+1}, L_{n+2}, F_{n+4}, 3 F_{n+2}\right) \text {; }  \tag{A4}\\
& \text { (ii) }\left(-3 L_{n-1}, L_{n-2}, L_{n+2}, 3 L_{n+1}\right) \text {; } \\
& \text { (iii) }\left(-3 F_{n-2},-F_{n-4}, L_{n-2}, 3 F_{n-1}\right) \text {. }
\end{align*}
$$

First, we verify that the above yield the desired solutions:

$$
\begin{align*}
& -2 F_{n-2}+F_{n}=-2 F_{n-2}+2 F_{n-2}+F_{n-3}=F_{n-3} ;  \tag{Al}\\
& F_{n-3}+F_{n}=F_{n-1}-F_{n-2}+F_{n-1}+F_{n-2}=2 F_{n-1} ; \\
& 2 F_{n-1}+F_{n}=F_{n-1}+F_{n+1}=L_{n} ; \\
& L_{n}+F_{n}=F_{n-1}+F_{n}+F_{n+1}=2 F_{n+1} ; \\
& 2 F_{n+1}+F_{n}=F_{n+1}+F_{n+2}=F_{n+3} \text {; } \\
& F_{n+3}+F_{n}=F_{n+2}+F_{n+1}+F_{n+2}-F_{n+1}=2 F_{n+2} \text {. Q.E.D. } \\
& -L_{n-1}+F_{n}=-F_{n-2}-F_{n}+F_{n}=-F_{n-2} ;-F_{n-2}+F_{n}=F_{n-1} ;  \tag{A2}\\
& F_{n-1}+F_{n}=F_{n+1} ; \quad F_{n+1}+F_{n}=F_{n+2} ; \quad F_{n+2}+F_{n}=L_{n+1} \cdot \quad \text { Q.E.D. } \\
& -4 L_{n-1}+5 F_{n}=-4 L_{n-1}+L_{n+1}+L_{n-1}=L_{n}+L_{n-1}-3 L_{n-1}  \tag{A3}\\
& =L_{n-1}+L_{n-2}-2 L_{n-1}=L_{n-2}-L_{n-1}=-L_{n-3} \text {; } \\
& -L_{n-3}+5 F_{n}=-L_{n-3}+L_{n+1}+L_{n-1}=-L_{n-1}+L_{n-2}+L_{n}+2 L_{n-1} \\
& =L_{n-2}+L_{n-1}+L_{n}=2 L_{n} \text {; } \\
& 2 L_{n}+5 F_{n}=2 L_{n}+L_{n-1}+L_{n+1}=L_{n}+L_{n-1}+L_{n}+L_{n+1} \\
& =L_{n+1}+L_{n+2}=L_{n+3} \text {; } \\
& L_{n+3}+5 F_{n}=L_{n+3}+L_{n-1}+L_{n+1}=L_{n+2}+L_{n+1}+L_{n+1}-L_{n}+L_{n+1} \\
& =L_{n+1}+L_{n}+3 L_{n+1}-L_{n}=4 L_{n+1} \text {. Q.E.D. } \\
& \text { (i) } 3 F_{n+1}+F_{n}=2 F_{n+1}+F_{n+2}=F_{n+1}+F_{n+3}=L_{n+2} \text {; }  \tag{A4}\\
& -3 F_{n-2}+F_{n}=-3 F_{n-3}-3 F_{n-4}+2 F_{n-2}+F_{n-3}  \tag{iii}\\
& =2 F_{n-2}-2 F_{n-3}-3 F_{n-4}=-F_{n-4}^{\prime} \text {; } \\
& -F_{n-4}+F_{n}=-F_{n-2}^{\prime}+F_{n-3}+F_{n-1}+F_{n-2}=L_{n-2} ;
\end{align*}
$$

$$
\begin{aligned}
L_{n-2}+F_{n} & =F_{n-3}+F_{n-1}+F_{n}=F_{n-1}-F_{n-2}+F_{n-1}+F_{n-1}+F_{n-2} \\
& =3 F_{n-1} \cdot \text { Q.E.D. }
\end{aligned}
$$

Although not required, it is informative to show how the preceding progressions were discovered. We illustrate the method for part (Al) of the problem. First, we note that the value 2 can be assumed only by the following seven admissible terms: $\left(F_{-3},-2 F_{-2}, 2 F_{-1}, L_{0}, 2 F_{1}, 2 F_{2}, F_{3}\right)$. If we suppose that these are special cases of the desired terms, not necessarily in proper order, we surmise that the general terms of the desired solution may be formed by adding $n$ to each suffix of the preceding list. If so, the asymptotic values of such terms are as follows, again, not necessarily in proper order:

$$
a^{n} 5^{-1 / 2} \cdot\left(\alpha^{-3},-2 \alpha^{-2}, 2 \alpha^{-1}, 5^{1 / 2}, 2 \alpha, 2 \alpha^{2}, a^{3}\right)
$$

The terms in parentheses may be crudely approximated as follows: (.24, -.76, $1.24,2.24,3.24,5.24,4.24)$. We now rearrange these last terms in ascending order of magnitude: ( $-.76, .24,1.24,2.24,3.24,4.24,5.24$ ), and note that all the terms are indeed in A.P. We now write down the terms of the first list corresponding to these last terms, as follows: $\left(-2 F_{-2}, F_{-3}, 2 F_{-1}, L_{0}, 2 F_{1}, F_{3}\right.$, $2 F_{2}$ ). Finally, we add $n$ to each suffix in this last septet, thereby forming the candidate for the desired general solution; as we have verified, this indeed generates the correct solution.

A similar process yields the solutions of the other parts of the problem, though in parts (A3) and (A4) the process is complicated by the fact that the choice of terms forming an A.P. is not unique; moreover, in (A4), a pair of "red herrings" occur, which cannot be used to form an A.P., but these are readily identifiable as such and may quickly be eliminated from consideration.
(B) The appropriate generalization is readily obtained by using the generalized Fibonacci and Lucas numbers defined as follows, for arbitrary constants $r$ and $s$ :

$$
U_{n}=r F_{n}+s F_{n-1}, \quad V_{n}=r L_{n}+s L_{n-1}, \text { for all integers } n .
$$

It is easy to see that the $U_{n}$ 's and $V_{n}$ 's satisfy the Fibonacci recurrence, but have different initial values, in general. From this, we see that the desired generalization is obtained by replacing $F$ by $U$ and $L$ by $V$ in (Al)-(A4); the differences in each A.P. will then be an appropriate multiple (either lor 5) of $U_{n}$, rather than of $E_{n}$. We illustrate only with case (A4)(i):

$$
\begin{aligned}
& \left(3 U_{n+1}, V_{n+2}, U_{n+4}, 3 U_{n+2}\right) \\
& =\left(3\left(r F_{n+1}+s F_{n}\right),\left(r L_{n+2}+s L_{n+1}\right),\left(r F_{n+4}+s F_{n+3}\right), 3\left(r F_{n+2}+s F_{n+1}\right)\right) \\
& =r\left(3 F_{n+1}, L_{n+2}, F_{n+4}, 3 F_{n+2}\right)+s\left(3 F_{n}, L_{n+1}, F_{n+3}, 3 F_{n+1}\right) ;
\end{aligned}
$$

from (A4)(i), each quadruplet in parentheses is in A.P., with common difference $F_{n}$ and $F_{n-1}$, respectively. Due to linearity, the general terms are also in A.P., with common difference $=r F_{n}+s F_{n-1}=U_{n}$. Q.E.D.

Also solved by L. Kuipers and the proposer.

## A Golden Result

H-423 Proposed by Stanley Rabinowitz, Littleton, MA
(Vol. 26, no. 3, August 1988)
Prove that each root of the equation

$$
F_{n} x^{n}+F_{n+1} x^{n-1}+F_{n+2} x^{n-2}+\cdots+F_{2 n-1} x+F_{2 n}=0
$$

has an absolute value near $\phi$, the golden ratio.
Solution by Tad White, University of California, Los Angeles, CA

Problem: Show that the zeros of the polynomial $F_{n} z^{n}+\cdots+F_{2 n}$ lie near the circle $|z|=\alpha$, where $\alpha$ is a positive root of $z^{2}-z-1=0$.
Solution: First divide through by $F_{n}$ to obtain a monic polynomial; we will examine the roots of

$$
f_{n}(z)=z^{n}+\frac{F_{n+1}}{F_{n}} z^{n-1}+\cdots+\frac{F_{2 n}}{F_{n}} .
$$

The following lemma gives us information about the coefficients of $f_{n}$.
Lemma 1: If $\beta$ is the negative root of $z^{2}-z-1=0$, then

$$
\left|\frac{F_{n+k}}{F_{n}}-\alpha^{k}\right| \leq \frac{\left|\beta^{n}\right| F_{k}}{F_{n}} \text {, for all } n, k \text {. }
$$

Proof: Using Binet's formula for $F_{n}$, we can write

$$
\sqrt{5}\left(F_{n+k}-F_{n} \alpha^{k}\right)=\left(\alpha^{n+k}-\beta^{n+k}\right)-\alpha^{k}\left(\alpha^{n}-\beta^{n}\right)=\beta^{n}\left(\beta^{k}-\alpha^{k}\right)=-\sqrt{5} \beta^{n} F_{k} ;
$$

dividing by $\sqrt{5} F_{n}$ and taking absolute values completes the proof. $\square$
If we define $g_{n}(z)=z^{n}+\alpha z^{n-1}+\cdots+\alpha^{n}$, then Lemma 1 tells us that the coefficients of $f_{n}$ and $g_{n}$ are close. To make this precise, we can define a norm on the vector space $P_{n}$ of complex polynomials of degree $\leq n$ via

$$
\|f\|=\sum_{k=0}^{n}\left|\alpha_{k}\right| \text { if } f(z)=\sum_{k=0}^{n} \alpha_{k} z^{k}
$$

Then Lemma 1 says that

$$
\left\|f_{n}-g_{n}\right\| \leq \beta^{n} \sum_{k=0}^{n} \frac{F_{k}}{F_{n}}<\beta^{n} \frac{F_{n+2}}{F_{n}}<3 \beta^{n} .
$$

Since $|\beta|<1$, this says $\left\|f_{n}-g_{n}\right\| \rightarrow 0$ as $n \rightarrow \infty$. Note also that

$$
g_{n}(z)=\frac{z^{n+1}-\alpha^{n+1}}{z-\alpha}
$$

so the roots of $g_{n}$ lie on the circle $|z|=\alpha$. Hence, we need only show that the locations of the zeros of a polynomial vary in some sense continuously with the coefficients. This can be made precise via the following lemma.

Lemma 2: Given a sufficiently small $\varepsilon>0$ and $f_{1} \in P_{n}$, there exists $\delta>0$ such that if $f_{0}$ is an element of $P_{n}$ with $\left\|f_{0}-f_{1}\right\|<\delta$, there exists a one-to-one correspondence between the roots $\zeta_{i}$ of $f_{0}$ and the roots $\eta_{i}$ of $f_{1}$ such that $\left|\zeta_{i}-\eta_{i}\right|<\varepsilon$ for each $i$.
Proof: Let $f_{t}=(1-t) f_{0}+t f_{1}$ for $0 \leq t \leq 1$; note that $f_{t} \in P_{n}$ for each $t$. Since the set of zeros of $f_{1}$ is discrete, and since $\varepsilon$ is small, $f_{1}$ does not vanish in the closed punctured ball of radius $\varepsilon$ around $\eta_{i}$. Observe that the evaluation maps $e_{z}: P_{n} \rightarrow \mathbb{C}$, given by $e_{z}(f)=f(z)$ are continuous with respect to the norm $\|\cdot\|$, and in fact uniformly continuous if we restrict $z$ to the compact set $|z|<2$ (which contains all of the roots $\eta_{i}$ in its interior). Therefore, we can choose $\delta$ such that $\left\|f_{0}-f_{1}\right\|<\delta$ implies that $f_{0}$ does not vanish on $\partial B\left(\eta_{i}, \varepsilon\right)$, where $B\left(\eta_{i}, \varepsilon\right)$ is the closed ball of radius $\varepsilon$ at $\eta_{i}$. Since $\left\|f_{0}-f_{1}\right\|$ is a monotonic function of $t$, we have that no $f_{t}$ vanishes on $\partial B\left(\eta_{i}, \varepsilon\right)$.

Assume further that $\varepsilon$ is small enough that the paths $\partial B\left(\eta_{i}, \varepsilon\right)$ are disjoint; then define the functions

$$
Z_{i}(t)=\frac{1}{2 \pi i} \int_{\partial B\left(n_{i}, \varepsilon\right)} \frac{f_{t},(z)}{f_{t}(z)} d z ;
$$

(please turn to page 282)

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## BOOKS AVAILABLE THROUGH THE FIBONACCI ASSOCIATION

Introduction to Fibonacci Discovery by Brother Alfred Brousseau. Fibonacci Association (FA), 1965.

Fibonacci and Lucas Numbers by Verner E. Hoggatt, Jr. FA, 1972.
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A Collection of Manuscripts Related to the Fibonacci Sequence-18th Anniversary Volume. Edited by Verner E. Hoggatt, Jr. and Marjorie Bicknell-Johnson. FA, 1980.

Fibonacci Numbers and Their Applications. Edited by A.N. Philippou, G.E. Bergum and A.F. Horadam.

Applications of Fibonacci Numbers. Edited by A.N. Philippou, A.F. Horadam and G.E. Bergum.

Please write to the Fibonacci Association, Santa Clara University, Santa Clara CA 95053, U.S.A., for current prices.


[^0]:    Note that the "purely periodic" hypothesis of the Theorem can be omitted if one substitutes asymptotic density for frequency, as the finite number of terms before $U_{n}$ becomes periodic modulo $M$ do not affect density. Our final result is well known but illustrates the Theorem's power.

    Corollary: Suppose that $U_{n}$ is u.d. (mod $\left.p^{k}\right)$ and is u.d. $(\bmod M)$, where $p$ is a prime that does not divide the length of the period of $U_{n}(\bmod M)$. "Then $U_{n}$ is u.d. (mod $\left.M \cdot p^{k}\right)$.

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