## C O M M U N I C A T I O N S

## Series A1: Mathematics and Statistics

## C O M M U N I C A T I O N S

FACULTY OF SCIENCES
DE LA FACULTE DES SCIENCES
UNIVERSITY OF ANKARA
DE L'UNIVERSITE D'ANKARA

Series A1: Mathematics and Statistics
Volume: 73
Number: 1
Year: 2024
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| Correspondence Address: | Print: |
| :--- | :--- |
| COMMUNICATIONS EDITORIAL OFFICE | Ankara University Press |
| Ankara University, Faculty of Sciences, | İncitaş Sokak No:10 06510 Beşevler |
| 06100 Tandoğan, ANKARA - TURKEY | ANKARA - TURKEY |
| Tel: (90) 312-2126720 Fax: (90) 312-2235000 | Tel: (90) 312-2136655 |
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## C O M M U N I C A T I O N

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# FRACTIONAL DIRAC SYSTEMS WITH MITTAG-LEFFLER KERNEL 

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#### Abstract

In this paper, we study some fractional Dirac-type systems with the Mittag-Leffler kernel. We extend the basic spectral properties of the ordinary Dirac system to the Dirac-type systems with the Mittag-Leffler kernel. First, this problem was handled in a continuous form. The self-adjointness of the operator produced by this system, the reality of its eigenvalues, and the orthogonality of the eigenfunctions have been investigated. Later, similar results were obtained by considering the discrete state.


## 1. Introduction

In recent years, the subject of fractional differential equations has become very popular among mathematicians. The investigation of all kinds of problems in the theory of differential equations under the framework of fractional has revealed a very wide field of study. The Dirac equation, which is one of the important equations in the history of physics, should also be investigated. Although fractional Sturm-Liouville problems have been investigated a lot, research on fractional Dirac equivalents is less. Contributing to the gap in this area in the literature is the main motivation of this research.

There are many types of fractional derivatives. One of them is the one based on the Mittag-Leffler function. Atangana and Baleanu introduced a new fractional derivative with the Mittag-Leffler kernel 4. In 1, Abdeljawad and Belanau defined integration with the part formula using the right fractional derivative and the right fractional integral corresponding to the Mittag-Leffler kernel. In 5, the

[^0]authors studied the discrete versions of these fractional derivatives. In 12, Mert et al. studied fractional Sturm-Liouville operators with the Mittag-Leffler kernels. With the help of the Laplace transform, Ercan is obtained the representation of solutions for fractional Dirac system with the Mittag-Leffler kernel ( |8). Yalçınkaya handled some Dirac systems with exponential kernel in 13. In 7, the authors studied a fractional Sturm-Liouville problem with exponential and Mittag-Leffler kernels.

In this study, we will investigate this type of fractional version of the Dirac system. Some basic features will be obtained for such systems. In the first chapter, the basic concepts and theorems that will be used in the study are given. In the following sections, the Dirac system with the Mittag-Leffler kernel in a continuous and discrete cases is discussed. This type of fractional Dirac system turns into the classical Dirac system by taking $\alpha \rightarrow 1$. It is transformed into a Riemann-Liouville type fractional Dirac system with a Laplace transform method. In this way, we examine these two systems under a single system. According to the knowledge of the authors, since there is no study on this subject in the literature, it will contribute to researchers working on this subject.

## 2. Preliminaries

This section covers the definitions and properties of fractional derivatives with the Mittag-Leffler kernel.

Definition 1. ( [1]) Let $u \in H^{1}(a, b)$ (the usual Sobolev space ), $a<b, \alpha \in[0,1]$. Then the definition of the left Caputo fractional derivative with the Mittag-Leffler kernel is given by

$$
\begin{equation*}
{ }_{a}^{A B C} D^{\alpha} u(\xi)=\frac{B(\alpha)}{1-\alpha} \int_{a}^{\xi} E_{\alpha}\left(\frac{-\alpha}{1-\alpha}(\xi-t)^{\alpha}\right) d(u(t)) \tag{1}
\end{equation*}
$$

where $B(\alpha)>0$ is a normalization function with $B(0)=B(1)=1$;

$$
\begin{equation*}
E_{\alpha, \beta}(t)=\sum_{k=0}^{\infty} \frac{t^{k}}{\Gamma(\alpha k+\beta)}, \tag{2}
\end{equation*}
$$

and $E_{\alpha}(t)=E_{\alpha, 1}(t)$. The convergence condition of infinite series (2) is $\operatorname{Re} \alpha>0$ and $\operatorname{Re} \beta>0$ ( 9$]$ ). Similarly, the left Riemann-Liouville fractional derivative with the Mittag-Leffler kernel has the following form

$$
\begin{equation*}
{ }_{a}^{A B R} D^{\alpha} u(\xi)=\frac{B(\alpha)}{1-\alpha} \frac{d}{d \xi} \int_{a}^{\xi} E_{\alpha}\left(\frac{-\alpha}{1-\alpha}(\xi-t)^{\alpha}\right) u(t) d t \tag{3}
\end{equation*}
$$

The associated fractional integral is given by

$$
{ }_{a}^{A B} I^{\alpha} u(\xi)=\frac{1-\alpha}{B(\alpha)} u(\xi)+\frac{\alpha}{B(\alpha) \Gamma(\alpha)} \int_{a}^{\xi}(\xi-t)^{\alpha-1} u(t) d t
$$

The right Caputo fractional derivative with the Mittag-Leffler kernel is given by

$$
\begin{equation*}
{ }^{A B C} D_{b}^{\alpha} u(\xi)=-\frac{B(\alpha)}{1-\alpha} \int_{\xi}^{b} E_{\alpha}\left(\frac{-\alpha}{1-\alpha}(t-\xi)^{\alpha}\right) u(t) d t \tag{4}
\end{equation*}
$$

and the right Riemann-Liouville derivative with the Mittag-Leffler kernel is defined by the formula

$$
{ }^{A B R} D_{b}^{\alpha} u(\xi)=-\frac{B(\alpha)}{1-\alpha} \frac{d}{d \xi} \int_{\xi}^{b} E_{\alpha}\left(\frac{-\alpha}{1-\alpha}(t-\xi)^{\alpha}\right) u(t) d t
$$

Moreover, the corresponding fractional integral is given by

$$
{ }^{A B} I_{b}^{\alpha} u(\xi)=\frac{1-\alpha}{B(\alpha)} u(\xi)+\frac{\alpha}{B(\alpha) \Gamma(\alpha)} \int_{\xi}^{b}(t-\xi)^{\alpha-1} u(t) d t .
$$

Proposition 1. ( $[1])$ Let $\alpha>0, p \geq 1, q \geq 1$, and $\frac{1}{p}+\frac{1}{q} \leq 1+\alpha(p \neq 1$ and $q \neq 1$ when $\left.\frac{1}{p}+\frac{1}{q}=1+\alpha\right)$.
(1) If $u \in L_{p}(a . b)$ and $v \in L_{q}(a . b)$, then

$$
\int_{a}^{b} u(\xi)_{a}^{A B} I^{\alpha} v(\xi) d \xi=\int_{a}^{b} v(\xi)^{A B} I_{b}^{\alpha} u(\xi) d \xi
$$

(2) If $u \in{ }^{A B} I_{b}^{\alpha}\left(L_{p}\right)$ and $v \in_{a}^{A B} I^{\alpha}\left(L_{q}\right)$, then

$$
\int_{a}^{b} u(\xi)_{a}^{A B R} D^{\alpha} v(\xi) d \xi=\int_{a}^{b} v(\xi)^{A B R} D_{b}^{\alpha} f(\xi) d \xi
$$

where

$$
{ }^{A B} I_{b}^{\alpha}\left(L_{p}\right)=\left\{u: u={ }^{A B} I_{b}^{\alpha} v, v \in L_{p}(a . b)\right\}
$$

and

$$
{ }_{a}^{A B} I^{\alpha}\left(L_{q}\right)=\left\{u: u={ }_{a}^{A B} I^{\alpha} v, v \in L_{q}(a . b)\right\} .
$$

Theorem 1. ( [1]) Let $u, v \in H^{1}(a, b), a<b$ and $\alpha \in(0,1)$. Then we have

$$
\begin{align*}
\int_{a}^{b} u(\xi)_{a}^{A B C} D^{\alpha} v(\xi) d \xi & =\int_{a}^{b} v(\xi)^{A B R} D_{b}^{\alpha} u(\xi) d \xi  \tag{1}\\
& +\left.\frac{B(\alpha)}{1-\alpha} v(\xi) E_{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} u(\xi)\right|_{a} ^{b}
\end{align*}
$$

where

$$
E_{\alpha, \beta, w, b^{-}} u(\xi)=\int_{\xi}^{b}(t-\xi)^{\beta-1} E_{\alpha, \beta}\left(w(t-\xi)^{\alpha}\right) u(t) d t, \xi<b
$$

2. 

$$
\begin{aligned}
\int_{a}^{b} u(\xi)^{A B C} D_{b}^{\alpha} v(\xi) d \xi & =\int_{a}^{b} v(\xi)_{a}^{A B R} D^{\alpha} u(\xi) d \xi \\
& -\frac{B(\alpha)}{1-\alpha} v(b) E_{\alpha, 1, \frac{-\alpha}{1-\alpha}, a^{+}} u(b) \\
& +\frac{B(\alpha)}{1-\alpha} v(a) E_{\alpha, 1, \frac{-\alpha}{1-\alpha}, a^{+}} u(a)
\end{aligned}
$$

where

$$
E_{\alpha, \beta, w, a+} u(\xi)=\int_{a}^{\xi}(\xi-t)^{\beta-1} E_{\alpha, \beta}\left(w(\xi-t)^{\alpha}\right) u(t) d t, \xi>a
$$

Let

$$
\begin{gathered}
\mathbb{N}_{a}=\{a, a+1, a+2, \ldots\}, \\
{ }_{b} \mathbb{N}=\{\ldots, b-2, b-1, b\}, \\
\mathbb{N}_{a, b}=\{a, a+1, a+2, \ldots, b\},
\end{gathered}
$$

where $a, b \in \mathbb{R}$ and $b-a$ is a positive integer.
Definition 2. ( $5 \cdot[6,10])$ Let $u: \mathbb{N}_{a} \rightarrow \mathbb{R}$ and $\alpha \in(0,1 / 2)$. Then the nabla discrete left Caputo difference with the Mittag-Leffler kernel is defined by

$$
{ }_{a}^{A B C} \nabla^{\alpha} u(\xi)=\frac{B(\alpha)}{1-\alpha} \sum_{i=a+1}^{\xi} \nabla_{i} u(i) E_{\bar{\alpha}}\left(\frac{-\alpha}{1-\alpha}, \xi-\rho(i)\right), \xi \in \mathbb{N}_{a+1}
$$

and the left Riemannn-Liouville one by

$$
{ }_{a}^{A B R} \nabla^{\alpha} u(\xi)=\frac{B(\alpha)}{1-\alpha} \nabla_{\xi} \sum_{i=a+1}^{\xi} u(i) E_{\bar{\alpha}}\left(\frac{-\alpha}{1-\alpha}, \xi-\rho(i)\right), \xi \in \mathbb{N}_{a+1}
$$

where $\rho(i)=i-1$; and the discrete the Mittag-Leffler kernel is defined by the formula

$$
E_{\bar{\alpha}}(\lambda, z)=\sum_{i=0}^{\infty} \lambda^{i} \frac{z^{\overline{i \alpha}}}{\Gamma(i \alpha+1)},
$$

where $z^{\overline{i \alpha}}=\prod_{i=0}^{i \alpha-1}(t+i), z^{\overline{0}}=1, t \in \mathbb{R}$. Moreover, the associated fractional sum function

$$
{ }_{a}^{A B} \nabla^{-\alpha} u(\xi)=\frac{1-\alpha}{B(\alpha)} u(\xi)+\frac{\alpha}{B(\alpha)} \nabla_{a}^{-\alpha} u(\xi), \quad \xi \in \mathbb{N}_{a+1}
$$

where

$$
\nabla_{a}^{-\alpha} u(\xi)=\frac{1}{\Gamma(\alpha)} \sum_{i=a+1}^{\xi}(\xi-\rho(i))^{\overline{\alpha-1}} u(i), \xi \in \mathbb{N}_{a+1} \text { (see [2], 3]). }
$$

Definition 3. ([5]) Let $u:_{b} \mathbb{N} \rightarrow \mathbb{R}$ and $\alpha \in(0,1 / 2)$. Then the nabla discrete right Caputo difference with the Mittag-Leffler kernel is defined by

$$
{ }^{A B C} \nabla_{b}^{\alpha} u(\xi)=\frac{-B(\alpha)}{1-\alpha} \sum_{i=\xi}^{b-1} \Delta u(i) E_{\bar{\alpha}}\left(\frac{-\alpha}{1-\alpha}, i-\rho(\xi)\right), \xi \in_{b-1} \mathbb{N}
$$

and the right Reimann-Liouville one by

$$
A B R^{A} \nabla_{b}^{\alpha} u(\xi)=\frac{-B(\alpha)}{1-\alpha} \Delta_{\xi} \sum_{i=\xi}^{b-1} u(i) E_{\bar{\alpha}}\left(\frac{-\alpha}{1-\alpha}, i-\rho(\xi)\right), \xi \in_{b-1} \mathbb{N}
$$

Further, the associated fractional sum is defined by

$$
{ }^{A B} \nabla_{b}^{-\alpha} u(\xi)=\frac{1-\alpha}{B(\alpha)} u(\xi)+\frac{\alpha}{B(\alpha)} \nabla_{b}^{-\alpha} u(\xi), \xi \in_{b-1} \mathbb{N}
$$

where

$$
\left.\nabla_{b}^{-\alpha} u(\xi)=\frac{1}{\Gamma(\alpha)} \sum_{i=\xi}^{b-1}(i-\rho(\xi))^{\overline{\alpha-1}} u(i), \xi \in_{b-1} \mathbb{N} \text { (see [2, 3] }\right)
$$

Theorem 2. ([5]) Let $u, v: \mathbb{N}_{a, b} \rightarrow \mathbb{R}$ and $\alpha \in(0,1 / 2)$. Then we have

$$
\begin{aligned}
\sum_{\xi=a+1}^{b-1} v(\xi)_{a}^{A B} \nabla^{-\alpha} u(\xi) & =\sum_{\xi=a+1}^{b-1} u(\xi)^{A B} \nabla_{b}^{-\alpha} v(\xi), \\
\sum_{\xi=a+1}^{b-1} v(\xi)_{a}^{A B R} \nabla^{\alpha} u(\xi) & =\sum_{\xi=a+1}^{b-1} u(\xi)^{A B R} \nabla_{b}^{\alpha} v(\xi),
\end{aligned}
$$

and

$$
\begin{aligned}
\sum_{\xi=a+1}^{b-1} u(\xi)_{a}^{A B C} \nabla^{\alpha} v(\xi) & =\sum_{\xi=a+1}^{b-1} v(\xi-1)^{A B R} \nabla_{b}^{\alpha} u(\xi-1) \\
& +v(b-1) \frac{B(\alpha)}{1-\alpha} E_{\frac{1}{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} u(b-1) \\
& -v(a) \frac{B(\alpha)}{1-\alpha} E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} u(a),
\end{aligned}
$$

where

$$
E_{\overline{\rho, \mu}, w, b^{-}}^{1} u(\xi)=\sum_{a=\xi}^{b-1}(a-\rho(\xi))^{\overline{\mu-1}} E_{\overline{\rho, \mu}}(w, a-\rho(\xi)) u(\xi), \xi \in_{b} \mathbb{N} .
$$

## 3. The Continuous Case

Let us consider the below continuous fractional Dirac system

$$
\begin{equation*}
L u:=B u+Q u=\lambda u, a \leq x \leq b<\infty \tag{5}
\end{equation*}
$$

where

$$
Q=\left(\begin{array}{cc}
p & 0 \\
0 & r
\end{array}\right), u:=\binom{u_{1}}{u_{2}}, B=\left(\begin{array}{cc}
0 & { }_{a}^{A B C} D^{\alpha} \\
A B R D_{b}^{\alpha} & 0
\end{array}\right)
$$

$\alpha \in(0,1), \lambda \in \mathbb{C} ; p, r \in C[a, b] ; p(x)>0, r(x)>0, \forall x \in[a, b]$. We also consider the following boundary conditions

$$
\begin{align*}
& \varkappa_{11} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} u_{1}(a)+\varkappa_{12} u_{2}(a)=0,  \tag{6}\\
& \varkappa_{21} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} u_{1}(b)+\varkappa_{22} u_{2}(b)=0, \tag{7}
\end{align*}
$$

with $\varkappa_{11}^{2}+\varkappa_{12}^{2} \neq 0$ and $\varkappa_{21}^{2}+\varkappa_{22}^{2} \neq 0$.
Now let's define the inner product suitable for this system. Let $L^{2}\left((a, b) ; \mathbb{R}^{2}\right)$ denotes the Hilbert space with the following inner product

$$
\begin{equation*}
(u, v):=\int_{a}^{b} u_{1} v_{1} d x+\int_{a}^{b} u_{2} v_{2} d x \tag{8}
\end{equation*}
$$

where

$$
u=\binom{u_{1}}{u_{2}}, v=\binom{v_{1}}{v_{2}}
$$

$u_{i}$ and $v_{i}(i=1,2)$ are real-valued continuous functions defined on $[a, b]$.
Theorem 3. The operator $L$ defined by (5)-(7) is formally self-adjoint on $L^{2}\left((a, b) ; \mathbb{R}^{2}\right)$.
Proof. Using (8), we get

$$
\begin{aligned}
(L u, v)-(u, L v) & =\int_{a}^{b}\left({ }_{a}^{A B C} D^{\alpha} u_{2}+p(x) u_{1}\right) v_{1} d x \\
& +\int_{a}^{b}\left({ }^{A B R} D_{b}^{\alpha} u_{1}+r(x) u_{2}\right) v_{2} d x \\
& -\int_{a}^{b} u_{1}\left({ }_{a}^{A B C} D^{\alpha} v_{2}+p(x) v_{1}\right) d x
\end{aligned}
$$

$$
\begin{aligned}
& -\int_{a}^{b} u_{2}\left({ }^{A B R} D_{b}^{\alpha} v_{1}+r(x) v_{2}\right) d x \\
& =\int_{a}^{b}\left({ }_{a}^{A B C} D^{\alpha} u_{2} v_{1}\right) d x+\int_{a}^{b} A B R D_{b}^{\alpha} u_{1} v_{2} d x \\
& -\int_{a}^{b} u_{1}\left({ }_{a}^{A B C} D^{\alpha} v_{2}\right) d x-\int_{a}^{b} u_{2}\left({ }^{A B R} D_{b}^{\alpha} v_{1}\right) d x
\end{aligned}
$$

where $u, v \in L^{2}\left((a, b) ; \mathbb{R}^{2}\right)$. From Proposition 1 and Theorem 1, we obtain

$$
\begin{equation*}
(L u, v)-(u, L v)=[u, v]_{b}-[u, v]_{a} \tag{9}
\end{equation*}
$$

where

$$
[u, v]_{x}=v_{2}(x) \frac{B(\alpha)}{1-\alpha} E_{\frac{1}{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}} u_{1}(x)-u_{2}(x) \frac{B(\alpha)}{1-\alpha} E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} v_{1}(x) .
$$

By conditions (6)-(7), we get the desired result.
Corollary 1. The eigenvalues of Eq. (5) subject to the boundary conditions (6)-(7) are real. The eigenfunctions corresponding to different eigenvalues of the system (5)-(7) are orthogonal.

Let us define the Wronskian of $u$ and $v$ by
$W(u, v)(x)=\left(\frac{B(\alpha)}{1-\alpha} E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} u_{1}(x)\right) v_{2}(x)-\left(\frac{B(\alpha)}{1-\alpha} E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} v_{1}(x)\right) u_{2}(x)$,
where

$$
u=\binom{u_{1}}{u_{2}}, v=\binom{v_{1}}{v_{2}} \in L^{2}\left((a, b) ; \mathbb{R}^{2}\right)
$$

Theorem 4. Let $v_{1}$ and $v_{2}$ be two solutions of Eq. (5). Then $W\left(v_{1}, v_{2}\right)$ is independent of $x$.

Proof. By (9), we obtain

$$
\left(\lambda v_{1}, v_{2}\right)-\left(v_{1}, \lambda v_{2}\right)=\left[v_{1}, v_{2}\right]_{b}-\left[v_{1}, v_{2}\right]_{a}
$$

since $L v_{1}=\lambda v_{1}$ and $L v_{2}=\lambda v_{2}$. Hence

$$
\left[v_{1}, v_{2}\right]_{b}=\left[v_{1}, v_{2}\right]_{a}=W\left(v_{1}, v_{2}\right)(a)
$$

Theorem 5. Any two solutions of the Eq. (5) are linearly dependent if and only if their Wronskian is zero.

Proof. Assume $v_{1}$ and $v_{2}$ be two linearly dependent solutions of Eq. (5). Then there exists a constant $\eta>0$ such that $v_{1}=\eta v_{2}$. Hence

$$
\begin{aligned}
W\left(v_{1}, v_{2}\right)(x) & =\left|\begin{array}{cc}
v_{11}(x) & \frac{B(\alpha)}{1-\alpha} E_{\alpha, 1}^{1} \frac{-\alpha}{1-\alpha}, b^{-} \\
v_{21}(x) & \frac{B(\alpha)}{1-\alpha} E_{12}(x) \\
\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-} \\
v_{22}(x)
\end{array}\right| \\
& =\left|\begin{array}{cc}
\eta v_{21}(x) & \eta \frac{B(\alpha)}{1-\alpha} E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} v_{22}(x) \\
v_{21}(x) & \frac{B(\alpha)}{1-\alpha} E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}} v_{22}(x)
\end{array}\right|=0 .
\end{aligned}
$$

On the other hand, if the Wronskian $W\left(v_{1}, v_{2}\right)(x)$ is zero for some $x$ in $[a, b]$, then we obtain

$$
v_{1}=\eta v_{2}
$$

i.e., $v_{1}$ and $v_{2}$ are linearly dependent on $[a, b]$.

Let us now give an example to illustrate our results.
Example 1. If we take $\alpha \rightarrow 1^{-}$in (5), we obtain the ordinary Dirac system ([11) defined as

$$
\left(\begin{array}{cc}
0 & 1 \\
1 & 0
\end{array}\right) \frac{d u}{d x}+Q u=\lambda u, a \leq x \leq b<\infty
$$

where

$$
Q=\left(\begin{array}{cc}
p & 0 \\
0 & r
\end{array}\right) \text { and } u:=\binom{u_{1}}{u_{2}}
$$

In fact, for $\alpha \in(0,1]$, the $A B R$ and $A B C$ fractional operators become well-defined due to the Mittag-Leffler kernel (2) doesn't have a convergence problem.

## 4. The Discrete Case

Let us consider the nabla discrete fractional Dirac systems

$$
\begin{gather*}
L_{1} u=C u+Q u=\lambda u, x \in \mathbb{N}_{a, b-1}  \tag{10}\\
Q=\left(\begin{array}{cc}
p & 0 \\
0 & r
\end{array}\right), u:=\binom{u_{1}}{u_{2}}, C=\left(\begin{array}{cc}
0 & { }_{a}^{A B C} \nabla^{\alpha} \\
A B R_{\nabla_{b}^{\alpha}} & 0
\end{array}\right),
\end{gather*}
$$

where $\alpha \in(0,1 / 2), \lambda \in \mathbb{C} ; p$ and $r$ are real-valued functions on $\mathbb{N}_{a, b-1} ; p(x)>$ $0, r(x)>0, \forall x \in \mathbb{N}_{a, b-1}$. We consider the following conditions

$$
\begin{align*}
& \varkappa_{11}\left({ }^{A B R^{\nabla}}{ }_{b}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}}\right) u_{1}(a)+\varkappa_{12} u_{2}(a)=0,  \tag{11}\\
& \varkappa_{21}\left({ }^{A B R} \nabla_{b}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}}\right) u_{1}(b-1)+\varkappa_{22} u_{2}(b-1)=0, \tag{12}
\end{align*}
$$

where $\varkappa_{11}^{2}+\varkappa_{12}^{2} \neq 0$ and $\varkappa_{21}^{2}+\varkappa_{22}^{2} \neq 0$.

Let $L_{\nabla}^{2}\left(\mathbb{N}_{a, b-1} ; \mathbb{R}^{2}\right)$ denotes the Hilbert space with the following inner product

$$
\langle u, v\rangle:=\sum_{x=a+1}^{b-1} u_{1}(x) v_{1}(x)+\sum_{x=a+1}^{b-1} u_{2}(x) v_{2}(x),
$$

where

$$
u=\binom{u_{1}}{u_{2}}, v=\binom{v_{1}}{v_{2}}
$$

$u_{i}$ and $v_{i}(i=1,2)$ are real-valued functions defined on $\mathbb{N}_{a, b-1}$.
Theorem 6. The operator $L_{1}$ defined by (10)-(12) is formally self-adjoint on $L_{\nabla}^{2}\left(\mathbb{N}_{a, b-1} ; \mathbb{R}^{2}\right)$.

Proof. Let $u, v \in L_{\nabla}^{2}\left(\mathbb{N}_{a, b-1} ; \mathbb{R}^{2}\right)$. Then we see that

$$
\begin{aligned}
& \left\langle L_{1} u, v\right\rangle-\left\langle u, L_{1} v\right\rangle=\sum_{x=a+1}^{b-1}\left({ }_{a}^{A B C} \nabla^{\alpha} u_{2}+p(x) u_{1}\right) v_{1}+\sum_{x=a+1}^{b-1}\left({ }^{A B R} \nabla_{b}^{\alpha} u_{1}+r(x) u_{2}\right) v_{2} \\
& -\sum_{x=a+1}^{b-1} u_{1}\left({ }_{a}^{A B C} \nabla^{\alpha} v_{2}+p(x) v_{1}\right)-\sum_{x=a+1}^{b-1} u_{2}\left({ }^{A B R} \nabla_{b}^{\alpha} v_{1}+r(x) v_{2}\right) \\
& =\sum_{x=a+1}^{b-1}{ }_{a}^{A B C} \nabla^{\alpha} u_{2} v_{1}+\sum_{x=a+1}^{b-1} p(x) u_{1}(x) v_{1}(x)+\sum_{x=a+1}^{b-1} A B R \nabla_{b}^{\alpha} u_{1} v_{2} \\
& +\sum_{x=a+1}^{b-1} r(x) u_{2}(x) v_{2}(x)-\sum_{x=a+1}^{b-1} u_{1}\left(\begin{array}{l}
A B C \\
a
\end{array} \nabla^{\alpha} v_{2}\right)-\sum_{x=a+1}^{b-1} p(x) u_{1}(x) v_{1}(x) \\
& -\sum_{x=a+1}^{b-1} u_{2}\left({ }^{A B R} \nabla_{b}^{\alpha} v_{1}\right)-\sum_{x=a+1}^{b-1} r(x) u_{2}(x) v_{2}(x) \\
& =\sum_{x=a+1}^{b-1}{ }_{a}^{A B C} \nabla^{\alpha} u_{2} v_{1}+\sum_{x=a+1}^{b-1} A B R \nabla_{b}^{\alpha} u_{1} v_{2} \\
& -\sum_{x=a+1}^{b-1} u_{1}\left({ }_{a}^{A B C} \nabla^{\alpha} v_{2}\right)-\sum_{x=a+1}^{b-1} u_{2}\left({ }^{A B R} \nabla_{b}^{\alpha} v_{1}\right) \\
& =\sum_{x=a+1}^{b-1} u_{2}(x-1)^{A B R} \nabla_{b}^{\alpha} v_{1}(x-1)+\left.\frac{B(\alpha)}{1-\alpha} u_{2}(x) E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} v_{1}(x)\right|_{a} ^{b-1} \\
& +\sum_{x=a+1}^{b-1}\left({ }^{A B R} \nabla_{b}^{\alpha} u_{1}\right) v_{2}-\sum_{x=a+1}^{b-1} v_{2}(x-1)^{C F R} \nabla_{b}^{\alpha} u_{1}(x-1) \\
& -\left.\frac{B(\alpha)}{1-\alpha} v_{2}(x) E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} u_{1}(x)\right|_{a} ^{b-1}-\sum_{x=a+1}^{b-1} u_{2}\left({ }^{A B R} \nabla_{b}^{\alpha} v_{1}\right)
\end{aligned}
$$

$$
\begin{aligned}
& =u_{2}(a)^{A B R} \nabla_{b}^{\alpha} v_{1}(a)-u_{2}(b-1)^{A B R} \nabla_{b}^{\alpha} v_{1}(b-1) \\
& +\frac{B(\alpha)}{1-\alpha} u_{2}(b-1) E_{\overline{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1} v_{1}(b-1)-\frac{B(\alpha)}{1-\alpha} u_{2}(a) E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} v_{1}(a) \\
& +v_{2}(b-1)^{A B R} \nabla_{b}^{\alpha} u_{1}(b-1)-v_{2}(a)^{A B R} \nabla_{b}^{\alpha} u_{1}(a) \\
& +\frac{B(\alpha)}{1-\alpha} v_{2}(a) E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} u_{1}(a)-\frac{B(\alpha)}{1-\alpha} v_{2}(b-1) E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} u_{1}(b-1) \\
& =z_{2}(b-1)\left(C F \nabla_{\nabla_{b}}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}}\right) y_{1}(b-1) \\
& -y_{2}(b-1)\left(C F R_{\nabla_{b}^{\alpha}}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E_{\frac{1}{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1-1}\right) z_{1}(b-1) \\
& -\left[\quad v_{2}(a)\left(A B R_{b}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}}\right) u_{1}(a)\right] \\
& \left.-u_{2}(a)\left(A B R_{\nabla_{b}^{\alpha}}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E_{\frac{1}{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}}^{1}\right) v_{1}(a)\right]
\end{aligned}
$$

It follows from (11) and (12) that

$$
\left\langle L_{1} u, v\right\rangle-\left\langle u, L_{1} v\right\rangle=0
$$

Corollary 2. All eigenvalues of the problem (10)-12) are real. Eigenfunctions corresponding to different eigenvalues are orthogonal.

Theorem 7. Let

$$
\left.W(u, v)(x)=\left\lvert\, \begin{array}{l}
A B R_{\nabla_{b}^{\alpha}}-\frac{B(\alpha)}{1-\alpha} E \frac{1}{\alpha, 1, \frac{-\alpha}{1-\alpha}, b^{-}} \\
A B R_{\nabla_{b}^{\alpha}}^{\alpha}-\frac{B(\alpha)}{1-\alpha} E \frac{1}{\alpha, 1}, \frac{-\alpha}{1-\alpha}, b^{-}
\end{array}\right.\right) u_{1}(x) \quad u_{2}(x) \mid
$$

where

$$
u=\binom{u_{1}}{u_{2}}, v=\binom{v_{1}}{v_{2}} \in L_{\nabla}^{2}\left(\mathbb{N}_{a, b-1} ; \mathbb{R}^{2}\right)
$$

and let $\theta_{1}$ and $\theta_{2}$ be two solutions of Eq. (5). Then $W\left(\theta_{1}, \theta_{2}\right)$ is independent of $x$. Moreover, any two linearly independent solutions $\varphi_{1}, \varphi_{2}$ of Eq. 10) are linearly dependent if and only if $W\left(\varphi_{1}, \varphi_{2}\right)=0$.

Proof. The proof is as in Theorem 4 and Theorem 5.

## 5. Conclusion

In this work, we have considered some fractional Dirac systems with MittagLeffler kernel. Firstly, a continuous fractional Dirac system with Mittag-Leffler kernel is studied. Its spectral properties are investigated. Later, the nabla discrete fractional Dirac system with Mittag-Leffler kernel is constructed. Similar properties are studied. Since Dirac systems have an important place in quantum physics, the properties of such systems are studied intensively. In this context, investigating fractional Dirac systems with Mittag-Leffler kernel will contribute to researchers working in this field. In the future, Green's function can be created for this system and eigenfunction expansions can be investigated.

Author Contribution Statements All authors contributed equally and significantly in writing this article. All authors read and approved the final manuscript.

Declaration of Competing Interests This work does not have any conflict of interest.

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# ON $\theta$-CONVEX CONTRACTIVE MAPPINGS WITH APPLICATION TO INTEGRAL EQUATIONS 

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#### Abstract

The fundamental goal of our paper is to study $\theta$-convex contractive mappings in metric spaces. We demonstrate some fixed point results for such mappings. Also, we give an application to integral equations of our results. Consequently, our results encompass numerous generalizations of the Banach contraction principle on metric space.


## 1. Introduction and Preliminaries

Banach 11 initially gave the Banach contraction principle which is an outstanding result in fixed point theory. Due to its significance, over the years, abounding researchers extended and generalized this contraction in many ways.

The notion of almost contraction was introduced by Berinde 2 . Also almost contraction was compared with other contractions and Berinde [2, [3, 4] demonstrated some fixed point theorems related to almost contraction.

Firstly Jleli 5 gave an attractive contraction called $\theta$-contraction and researched the uniqueness and existence of these mappings in complete metric spaces. After Jleli's first article 5, some different fixed point theorems were introduced Jleli 6, Hussain 7 and Imdad 8 by changing and relaxing the conditions of $\mho$.

In recent years, a remarkable generalization of the Banach contraction principle is the theorem by Istratescu 9. Again, Istarescu studied convex contractions in 9, 10, 11. Since Istratescu's fixed point theorems, many authors studied numerous generalizations and applications of the result of Istratescu (see $12-24$ ).

[^1]Ciric 25 used the concept of orbitally continuous for proving the uniqueness and existence of the fixed point mappings. Afterwards, Bisht 15 proved some fixed point theorems by replacing the continuity condition with orbital continuity.

Merging the ideas of Istratescu 9 and Jleli 5, we introduce a generalization of convex type contractions. The goal of our paper is to introduce generalized $\theta$-convex contractive mappings and to demonstrate some fixed point theorems. Theorems that have been demonstrated in our paper are generalizations of a variety of results in the literature.

Now, at first we mention some fundemental definitions and notions related to our work.
$F(h)=\{t \in W: h t=t\}$ is fixed point of $h$.
Bisht 15 gave the following definition instead of continuity condition to be used their theorems.

Definition 1. 15 Let $(W, \varrho)$ be a metric space and $h$ be a self mapping on $W$. We say that $h$ is orbitally continuous at a point $u \in W$ if $\lim _{j \rightarrow \infty} h^{n_{j}} t=u$ implies that $\lim _{j \rightarrow \infty} h^{n_{j}} t=h u$.

Berinde 2, 3, 4 gave the concepts of almost contraction, multivalued almost contraction and the continuity of almost contractions.

Definition 2. [2] Let $(W, \varrho)$ be a metric space and $h$ be a self mapping on $W . h$ is called an almost contraction if there exists a constant $\zeta \in(0,1)$ and $L \geq 0$ such that

$$
\varrho(h t, h s) \leq \zeta \varrho(t, s)+L \varrho(s, h t)
$$

for all $t, s \in W$.
Firstly, Jleli 5 gave the concept of $\theta$-contraction mappings and the following family.

Let $\mathcal{V}$ denotes the set of all mappings $\theta:(0, \infty) \rightarrow(1, \infty)$ which hold the following conditions:
(1) $\theta$ is strictly increasing;
(2) for all sequence $\left\{\eta_{n}\right\} \subset(0, \infty), \lim _{n \rightarrow \infty} \theta\left(\eta_{n}\right)=1$ if and only if $\lim _{n \rightarrow \infty} \eta_{n}=$ $0 ;$
(3) there exist $\ell \in(0, \infty]$ and $r \in(0,1)$ such that $\lim _{n \rightarrow \infty} \frac{\theta(\eta)-1}{(\eta)^{r}}=\ell$.
$\Upsilon$ be the set of nondecreasing functions $\varsigma:[0,+\infty) \rightarrow[0,+\infty)$ such that $\sum_{j=1}^{+\infty} \varsigma^{j}(\eta)<+\infty$ for each $\eta>0$, where $\varsigma^{j}$ is the $j-$ th iterate of $\varsigma$.

Remark 1. Each function $\varsigma \in \Upsilon$ satisfies $\lim _{n \rightarrow \infty} \varsigma^{n}(\eta)=0$ and $\varsigma(\eta)<\eta$ for all $\eta>0$.

Firstly, Jleli 5 gave the definition of $\theta$-contraction as follows.

Definition 3. Let $(W, \varrho)$ be a metric space and $h: W \rightarrow W$ be a self-mapping. Then $h$ is called $\theta$-contraction if there exist $\kappa \in(0,1)$ such that

$$
\theta(\varrho(h t, h s)) \leq[\theta(\varrho(t, s))]^{\kappa}
$$

for all $t, s \in W$, with $h t \neq h s$.
Istratescu 9, 10 gave the following definitions.
Definition 4. Let $(W, \varrho)$ be a metric space and $h: W \rightarrow W$ be a self-mapping. Then $h$ is called convex contraction of order 2 if there exist $d_{1}, d_{2} \in(0,1)$ such that $d_{1}+d_{2}<1$ and

$$
\varrho\left(h^{2} t, h^{2} s\right) \leq d_{1} \varrho(h t, h s)+d_{2} \varrho(t, s)
$$

for all $t, s \in W$.
Definition 5. Let $(W, \varrho)$ be a metric space and $h: W \rightarrow W$ be a self-mapping. Then $h$ is called two-sided convex contraction mappings if there exist $d_{1}, d_{2}, d_{3}, d_{4} \in$ $(0,1)$ such that $d_{1}+d_{2}+d_{3}+d_{4}<1$ and

$$
\varrho\left(h^{2} t, h^{2} s\right) \leq d_{1} \varrho(t, h t)+d_{2} \varrho\left(h t, h^{2} t\right)+d_{3} \varrho(s, h s)+d_{4} \varrho\left(h s, h^{2} s\right)
$$

for all $t, s \in W$.

## 2. Main Results

In this chapter, we give concept of generalized $\theta$-convex contractions in metric spaces. We demonstrate some fixed point results for such contractions on metric spaces. The following Theorem's hypothesis are basically weaker than the set of contraction type mappings.

Now, we will give the definition of generalized $\theta$-convex contractive mappings.
Definition 6. Let $(W, \varrho)$ be a metric space and $h: W \rightarrow W$ be a self-mapping. Then $h$ is called generalized $\theta$-convex contraction if there exist $L \geq 0, \varsigma \in \Upsilon$ and $\kappa \in(0,1)$ such that

$$
\begin{equation*}
\varrho\left(h^{2} t, h^{2} s\right)>0 \Rightarrow \theta\left(\varrho\left(h^{2} t, h^{2} s\right)\right) \leq\left[\theta\left(\varsigma\left(M_{I}(t, s)\right)\right)\right]^{\kappa}+L N_{I}(t, s) \tag{1}
\end{equation*}
$$

where $\theta \in \mathcal{V}$ and

$$
\begin{aligned}
& M_{I}(t, s)=\max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}, \\
& N_{I}(t, s)=\min \left\{\varrho(t, h t), \varrho(s, h s), \varrho(t, h s), \varrho(s, h t), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}, \\
& \text { for all } t, s \in W
\end{aligned}
$$

Remark 2. Every convex contraction of order 2 and two-sided convex contraction are a generalized $\theta$-convex contraction. Also, every $\theta$-contraction is a generalized $\theta$-convex contraction. But the reverse doesn't have to be true.

Since, our novel class of contractive type mappings is more general, it will be more advantageous to work using this new class.

The following theorem is our first result related to generalized $\theta$-convex contractive mappings.

Theorem 1. Let $(W, \varrho)$ be a complete metric space and $h: W \rightarrow W$ be a generalized $\theta$-convex contraction. If $h$ is either orbitally continuous on $W$ or $h$ is continuous, then $h$ has a unique fixed point.

Proof. Starting at the point $t_{0} \in W$, the sequence $\left\{t_{n}\right\}$ is constructed by $t_{n}=$ $h t_{n-1}=h^{n} t_{0}, n \geq 1$. If $t_{n_{0}+1}=t_{n_{0}}$ for any $n_{0} \in \mathbb{N} \cup\{0\}$, then it is clear that, $t_{n_{0}}$ is a fixed point of $h$. Consequently, assume that $t_{n_{0}+1} \neq t_{n_{0}}$ for all $n_{0} \in \mathbb{N} \cup\{0\}$. Setting $m=\max \left\{\varrho\left(t_{0}, t_{1}\right), \varrho\left(t_{1}, t_{2}\right)\right\}$. First of all, we show that $\left\{\varrho\left(t_{n}, t_{n+1}\right)\right\}$ is a strictly nonincreasing sequence in $W$. Since $h$ is a generalized $\theta$-convex contraction, using Remark 1 and from the first axiom of $\theta$, we have

$$
\begin{aligned}
\theta\left(\varrho\left(t_{2}, t_{3}\right)\right)= & \theta\left(\varrho\left(h^{2} t_{0}, h^{2} t_{1}\right)\right) \\
\leq & {\left[\theta\left(\varsigma\left(\max \left\{\begin{array}{c}
\varrho\left(t_{0}, t_{1}\right), \varrho\left(h t_{0}, h t_{1}\right), \varrho\left(t_{0}, h t_{0}\right), \\
\varrho\left(h t_{0}, h^{2} t_{0}\right), \varrho\left(t_{1}, h t_{1}\right), \varrho\left(h t_{1}, h^{2} t_{1}\right)
\end{array}\right\}\right)\right)\right]^{\kappa} } \\
& +L \min \left\{\begin{array}{c}
\varrho\left(t_{0}, h t_{0}\right), \varrho\left(t_{1}, h t_{1}\right), \varrho\left(t_{0}, h t_{1}\right), \\
\varrho\left(t_{1}, h t_{0}\right), \varrho\left(h t_{0}, h^{2} t_{0}\right), \varrho\left(h t_{1}, h^{2} t_{1}\right)
\end{array}\right\} \\
= & {\left[\theta\left(\varrho\left(\max \left\{\varrho\left(t_{0}, t_{1}\right), \varrho\left(t_{1}, t_{2}\right), \varrho\left(t_{2}, t_{3}\right)\right\}\right)\right)\right]^{\kappa} } \\
\leq & {\left[\theta\left(\max \left\{m, \varrho\left(t_{2}, t_{3}\right)\right\}\right)\right]^{\kappa} . }
\end{aligned}
$$

If $\max \left\{m, \varrho\left(t_{2}, t_{3}\right)\right\}=\varrho\left(t_{2}, t_{3}\right)$, then we have

$$
\theta\left(\varrho\left(t_{2}, t_{3}\right)\right) \leq\left[\theta\left(\varrho\left(t_{2}, t_{3}\right)\right)\right]^{\kappa}
$$

If we take $\ln$ two both sides of the inquality, then we have

$$
\ln \theta\left(\varrho\left(t_{2}, t_{3}\right)\right) \leq \kappa \ln \left[\theta\left(\varrho\left(t_{2}, t_{3}\right)\right)\right]
$$

which is a contradiction. Hence, we get

$$
\max \left\{m, \varrho\left(t_{2}, t_{3}\right)\right\}=m=\max \left\{\varrho\left(t_{0}, t_{1}\right), \varrho\left(t_{1}, t_{2}\right)\right\} .
$$

Since $\varsigma(\eta)<\eta$ for all $\eta>0$, we have

$$
\begin{aligned}
\theta\left(\varrho\left(t_{3}, t_{4}\right)\right) \leq & {\left[\theta\left(\varsigma\left(\max \left\{\begin{array}{c}
\varrho\left(t_{1}, t_{2}\right), \varrho\left(h t_{1}, h t_{2}\right), \varrho\left(t_{1}, h t_{1}\right), \\
\varrho\left(h t_{1}, h^{2} t_{1}\right), \varrho\left(t_{2}, h t_{2}\right), \varrho\left(h_{2}, h^{2} t_{2}\right)
\end{array}\right\}\right)\right)\right]^{\kappa} } \\
& +L \min \left\{\begin{array}{c}
\varrho\left(t_{1}, h t_{1}\right), \varrho\left(t_{2}, h t_{2}\right), \varrho\left(t_{1}, h t_{2}\right), \\
\varrho\left(t_{2}, h t_{1}\right), \varrho\left(h t_{1}, h^{2} t_{1}\right), \varrho\left(h t_{2}, h^{2} t_{2}\right)
\end{array}\right\} \\
\leq & {\left[\theta\left(\max \left\{\varrho\left(t_{1}, t_{2}\right), \varrho\left(t_{2}, t_{3}\right), \varrho\left(t_{3}, t_{4}\right)\right\}\right)\right]^{\kappa} . }
\end{aligned}
$$

If $\max \left\{\varrho\left(t_{1}, t_{2}\right), \varrho\left(t_{2}, t_{3}\right), \varrho\left(t_{3}, t_{4}\right)\right\}=\varrho\left(t_{3}, t_{4}\right)$, then we obtain

$$
\theta\left(\varrho\left(t_{3}, t_{4}\right)\right) \leq\left[\theta\left(\varrho\left(t_{3}, t_{4}\right)\right)\right]^{\kappa} .
$$

If we take $\ln$ two both sides of the inequality, then we have

$$
\ln \theta\left(\varrho\left(t_{3}, t_{4}\right)\right) \leq \kappa \ln \left[\theta\left(\varrho\left(t_{3}, t_{4}\right)\right)\right] .
$$

This is one more contradiction, from which it is concluded that max $\left\{\varrho\left(t_{1}, t_{2}\right), \varrho\left(t_{2}, t_{3}\right)\right\}>$ $\varrho\left(t_{3}, t_{4}\right)$. Thus, $m>\varrho\left(t_{2}, t_{3}\right)>\varrho\left(t_{3}, t_{4}\right)$. Hence, by induction one can get
$\left\{\varrho\left(t_{n}, t_{n+1}\right)\right\}$ is a strictly nonincreasing sequence in $W$. This implies that

$$
\begin{aligned}
\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq & {\left[\theta\left(\varsigma\left(\max \left\{\begin{array}{c}
\varrho\left(t_{n-2}, t_{n-1}\right), \varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n-2}, t_{n-1}\right), \\
\varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n}, t_{n+1}\right)
\end{array}\right\}\right)\right)\right]^{\kappa} } \\
& +L \min \left\{\begin{array}{c}
\varrho\left(t_{n-2}, t_{n-1}\right), \varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n-2}, t_{n}\right), \\
\varrho\left(t_{n-1}, t_{n-1}\right), \varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n}, t_{n+1}\right)
\end{array}\right\} \\
\leq & {\left[\theta\left(\max \left\{\varrho\left(t_{n-2}, t_{n-1}\right), \varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n}, t_{n+1}\right)\right\}\right)\right]^{\kappa} . }
\end{aligned}
$$

If $\max \left\{\varrho\left(t_{n-2}, t_{n-1}\right), \varrho\left(t_{n-1}, t_{n}\right), \varrho\left(t_{n}, t_{n+1}\right)\right\}=\varrho\left(t_{n}, t_{n+1}\right)$ then we get

$$
\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq\left[\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)\right]^{\kappa}
$$

which is once again contradiction. Therefore, we have

$$
\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq\left[\theta\left(\max \left\{\varrho\left(t_{n-2}, t_{n-1}\right), \varrho\left(t_{n-1}, t_{n}\right)\right\}\right)\right]^{\kappa}
$$

and

$$
\begin{aligned}
\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq & {\left[\theta\left(\varrho\left(t_{n-1}, t_{n}\right)\right)\right]^{\kappa} } \\
\leq & {\left[\theta\left(\varrho\left(t_{n-2}, t_{n-1}\right)\right)\right]^{\kappa^{2}} } \\
& \vdots \\
\leq & {[\theta(m)]^{\kappa^{l}}, }
\end{aligned}
$$

whenever $l=2 n$ or $l=2 n+1$, for $l \geq 1$. Hence, we have

$$
\begin{equation*}
1 \leq \theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq[\theta(m)]^{\kappa^{l}}, \text { for all } l \geq 1 \tag{2}
\end{equation*}
$$

Letting $n \rightarrow \infty$, following two cases arise.
Case 1. $1 \leq \theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq[\theta(m)]^{\kappa^{n}}$, for all $n \geq 2$ and $n$ is even.
Case 2. $1 \leq \theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right) \leq[\theta(m)]^{\kappa^{n-1}}$, for all $n \geq 3$ and $n$ is odd.
From Case 1 and Case 2 we get $\lim _{n \rightarrow \infty} \theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)=1$. By the second axiom of $\theta$, we get $\lim _{n \rightarrow \infty} \varrho\left(t_{n}, t_{n+1}\right)=0$. From the third axiom of $\theta$, there exist $\ell \in(0, \infty]$ and $r \in(0,1)$

$$
\lim _{n \rightarrow \infty} \frac{\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)-1}{\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r}}=\ell
$$

Assume that $\ell<\infty$ and $T=\frac{\ell}{2}>0$. From the limit definition, there exists $n_{0} \in \mathbb{N}$ such that

$$
\left|\frac{\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)-1}{\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r}}-\ell\right| \leq \top \text { for all } n \geq n_{0}
$$

which implies that

$$
\frac{\theta \varrho\left(t_{n}, t_{n+1}\right)-1}{\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r}} \leq \ell-\top=\top \text { for all } n \geq n_{0}
$$

Therefore, we have

$$
n\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r} \leq \mathbb{k} n\left[\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)-1\right] \text { for all } n \geq n_{0}
$$

where $\mathbb{k}=\frac{1}{\top}$. Assume that $T>0$ is an arbitrary number and $\ell=\infty$. From the limit definition, there exists $n_{0} \in \mathbb{N}$ such that

$$
\frac{\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)-1}{\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r}} \geq \top \text { for all } n \geq n_{0}
$$

which implies that

$$
\begin{equation*}
n\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r} \leq \mathbb{k} n\left[\theta\left(\varrho\left(t_{n}, t_{n+1}\right)\right)-1\right] \text { for all } n \geq n_{0} \tag{3}
\end{equation*}
$$

where $\mathbb{k}=\frac{1}{\top}$. Therefore, in two cases there exists $n \geq n_{0}$ and $\mathbb{k}>0$ such that (2.3) is satisfied. Using (2.2), we get

$$
n\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r} \leq \mathbb{k} n\left([\theta(m)]^{k^{l}}-1\right) \text { for all } l \geq 2 n_{0}+1 \text { or } l \geq 2 n_{0}
$$

Letting $n \rightarrow \infty$, we get $\lim _{n \rightarrow \infty} n\left[\varrho\left(t_{n}, t_{n+1}\right)\right]^{r}=0$. Hence, there exists $n_{1} \in \mathbb{N}$ such that

$$
\varrho\left(t_{n}, t_{n+1}\right) \leq \frac{1}{n^{\frac{1}{r}}} \text { for all } n \geq n_{1}
$$

Now, we will demostrate that $\left\{t_{n}\right\}$ is a Cauchy sequence. For all $p>q \geq n_{1}$, we get

$$
\begin{aligned}
\varrho\left(t_{p}, t_{q}\right) & \leq \varrho\left(t_{p}, t_{p-1}\right)+\varrho\left(t_{p-1}, t_{p-2}\right)+\cdots+\varrho\left(t_{q+1}, t_{q}\right) \\
& \leq \sum_{j=q}^{p-1} \varrho\left(t_{j}, t_{j+1}\right) \\
& <\sum_{j=q}^{\infty} \varrho\left(t_{j}, t_{j+1}\right) \\
& \leq \sum_{j=q}^{\infty} \frac{1}{j^{\frac{1}{r}}}
\end{aligned}
$$

Since $\sum_{j=q}^{\infty} \frac{1}{j^{\frac{1}{r}}}$ is convergent, $\lim _{p, q \rightarrow \infty} \varrho\left(t_{p}, t_{q}\right)=0$. Hence, we get that $\left\{t_{n}\right\}$ is a Cauchy sequence in $W$. Since $(W, \varrho)$ is a complete metric space, there exists $u \in W$ such that $t_{n} \rightarrow u$. Assume that $h$ is continuous. Since $t_{n} \rightarrow u \in W$ and $W$ is complete metric space, we get

$$
\varrho(u, h u)=\lim _{n \rightarrow \infty} \varrho\left(t_{n}, h t_{n}\right)=\lim _{n \rightarrow \infty} \varrho\left(t_{n}, t_{n+1}\right)=0
$$

Therefore $u \in F(h)$. Again, assume that $h$ is orbitally continuous on $W$, then

$$
t_{n+1}=h t_{n}=h\left(h^{n} t_{0}\right) \rightarrow h u \text { as } n \rightarrow \infty .
$$

Since $W$ is complete metric space, $h u=u$ that is $u \in F(h)$. Now, assume that $u$ and $v$ are arbitrary two fixed point of $h$. Then we get

$$
\theta(\varrho(u, v))=\theta\left(\varrho\left(h^{2} u, h^{2} v\right)\right) \leq\left[\theta\left(\varsigma\left(M_{I}(u, v)\right)\right)\right]^{\kappa}+L N_{I}(u, v)
$$

$$
\begin{aligned}
\leq & {\left[\theta\left(\varsigma\left(\max \left\{\begin{array}{c}
\varrho(u, v), \varrho(h u, h v), \varrho(u, h u), \varrho(v, h v), \\
\varrho\left(h u, h^{2} u\right), \varrho\left(h v, h^{2} v\right)
\end{array}\right\}\right)\right)\right]^{\kappa} } \\
& +L \min \left\{\begin{array}{c}
\varrho(u, h u), \varrho(v, h v), \varrho(u, h v), \varrho(v, h u), \\
\varrho\left(h u, h^{2} u\right), \varrho\left(h v, h^{2} v\right)
\end{array}\right\} \\
\leq & {[\theta(\varrho(u, v))]^{\kappa} . }
\end{aligned}
$$

Thus we get

$$
\theta(\varrho(u, v)) \leq[\theta(\varrho(u, v))]^{\kappa}
$$

If we take $\ln$ two both sides of the inquality, then we obtain

$$
\ln \theta(\varrho(u, v)) \leq \kappa \ln \theta(\varrho(u, v))
$$

Since $\kappa \in(0,1)$, it is a contradiction. Hence $u=v$, that is, $h$ has a unique fixed point in $W$.

Now, we shall give an example to illustrate the generality of Theorem 1
Example 1. Let $(W, \varrho)$ be a metric space, $h$ be a self mapping on $W$ and $\theta(t)=e^{\sqrt{t}}$ for $t>0$, that is, $\theta \in \mho$. Assume that $h$ is a convex contraction of type-2 for all $t, s \in W$ with $\varrho\left(h^{2} t, h^{2} s\right)>0, B=\sum_{j=1}^{6} d_{j}<1$ and $d_{j} \geq 0$ for all $j=1,2, \ldots, 6$.

$$
\begin{aligned}
\varrho\left(h^{2} t, h^{2} s\right) \leq & d_{1} \varrho(t, s)+d_{2} \varrho(h t, h s)+d_{3} \varrho(t, h t)+d_{4} \varrho(s, h s) \\
& +d_{5} \varrho\left(h t, h^{2} t\right)+d_{6} \varrho\left(h s, h^{2} s\right) \\
\leq & \sum_{j=1}^{6} d_{j} \max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\} \\
\leq & B \max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}
\end{aligned}
$$

where $t, s \in W$ with $\varrho\left(h^{2} t, h^{2} s\right)>0$. We obtain that

$$
\varrho\left(h^{2} t, h^{2} s\right) \leq B M_{I}(t, s)
$$

Taking $\varsigma(t)=B^{\frac{1}{2}} t$, we have

$$
e^{\sqrt{\rho\left(h^{2} t, h^{2} s\right)}} \leq e^{B^{\frac{1}{4}} \sqrt{M_{I}(t, s)}}=\left[e^{\sqrt{\varrho\left(M_{I}(t, s)\right)}}\right]^{\kappa}
$$

where $\kappa=B^{\frac{1}{4}}$. Since $\theta(t)=e^{\sqrt{t}}$ for $t>0$, we deduce that

$$
\begin{aligned}
\theta\left(\varrho\left(h^{2} t, h^{2} s\right)\right. & \leq\left[\theta\left(\varsigma\left(M_{I}(t, s)\right)\right)\right]^{\kappa} \\
& \leq\left[\theta\left(\varsigma\left(M_{I}(t, s)\right)\right)\right]^{\kappa}+L N_{I}(t, s)
\end{aligned}
$$

where $L \geq 0$. This shows that, $h$ is a generalized $\theta$-convex contractive mapping.
Remark 3. Above example show that our contraction condition generalizes Istratescu's contraction conditions [9], 10].

Definition 7. Let $(W, \varrho)$ be a metric space. A self-mapping $h: W \rightarrow W$ is called an almost $\theta$-convex contraction if there exist $L \geq 0$ and $\kappa \in(0,1)$ such that

$$
\varrho\left(h^{2} t, h^{2} s\right)>0 \Rightarrow \theta\left(\varrho\left(h^{2} t, h^{2} s\right) \leq\left[\theta\left(M_{I}(t, s)\right)\right]^{\kappa}+L N_{I}(t, s)\right.
$$

where $\theta \in \mathcal{V}$ and

$$
\begin{aligned}
M_{I}(t, s) & =\max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\} \\
N_{I}(t, s) & =\min \left\{\varrho(t, h t), \varrho(s, h s), \varrho(t, h s), \varrho(s, h t), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}
\end{aligned}
$$

for all $t, s \in W$.
Definition 6 and Definition 7 generalize and merge the results derived by Jleli 5 and Istratescu [9, 10, and some other connected results in the literature. Also, our novel contractions can be considered as an attracted generalization of Darbo's fixed point problem 26, 27.

Corollary 1. Let $(W, \varrho)$ be a complete metric space and $h: W \rightarrow W$ be an almost $\theta$-convex contraction. If $h$ is either orbitally continuous on $W$ or $h$ is continuous, then $h$ has a unique fixed point that is $u=h u, u \in W$.

If we take $L=0$ in Theorem 1, then we obtain the following corollary.
Corollary 2. Let $(W, \varrho)$ be a metric space and $h: W \rightarrow W$ be a self-mapping. If there exist $\varsigma \in \Upsilon$ and $\kappa \in(0,1)$ such that

$$
\begin{equation*}
\varrho\left(h^{2} t, h^{2} s\right)>0 \Rightarrow \theta\left(\varrho\left(h^{2} t, h^{2} s\right)\right) \leq\left[\theta\left(\varsigma\left(M_{I}(t, s)\right)\right)\right]^{\kappa} \tag{4}
\end{equation*}
$$

where $\theta \in \mathcal{U}$ and

$$
M_{I}(t, s)=\max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}
$$

for all $t, s \in W$. Also, assume that $h$ is either orbitally continuous on $W$ or $h$ is continuous, then $u=h u, u \in W$.

By taking $L=0$ and not considering $\varsigma \in \Upsilon$ in Theorem 1] we deduce the following corollary.

Corollary 3. Let $(W, \varrho)$ be a metric space and a self-mapping $h$ on $W$. If there exist $\kappa \in(0,1)$ such that

$$
\begin{equation*}
\varrho\left(h^{2} t, h^{2} s\right)>0 \Rightarrow \theta\left(\varrho\left(h^{2} t, h^{2} s\right)\right) \leq\left[\theta\left(M_{I}(t, s)\right)\right]^{\kappa} \tag{5}
\end{equation*}
$$

where $\theta \in \mho$ and

$$
M_{I}(t, s)=\max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}
$$

for all $t, s \in W$. Also, assume that $h$ is either orbitally continuous on $W$ or $h$ is continuous, then $u=h u, u \in W$.

We get the following results as shown in Example 1 .

Corollary 4. Let $(W, \varrho)$ be a metric space and a self-mapping $h$ on $W$. For all $t, s \in W$,

$$
\varrho\left(h^{2} t, h^{2} s\right) \leq B \max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\}
$$

where $B \in[0,1)$. Also, assume that $h$ is either orbitally continuous on $W$ or $h$ is continuous, then $u=h u, u \in W$.

Corollary 5. Let $(W, \varrho)$ be a metric space and $h$ is a convex contraction of type-2 on $W$. Also, assume that $h$ is either orbitally continuous on $W$ or $h$ is continuous, then $u=h u, u \in W$.

## 3. Application

Now, we give an application of our result for nonlinear integral equations.

$$
\begin{equation*}
t(u)=\vartheta(u)+\int_{e}^{f} K(u, v, t(v)) d v \tag{6}
\end{equation*}
$$

where $e, f \in \mathbb{R}, C[e, f]=\{h:[e, f] \rightarrow \mathbb{R}$ continuous functions $\}, t \in C([e, f], \mathbb{R}), K$ : $[e, f] \times[e, f] \times \mathbb{R} \rightarrow \mathbb{R}$ and $\vartheta:[e, f] \rightarrow \mathbb{R}$.

Theorem 2. Consider the integral equation (3.1). Assume that the following conditions satisfy:
(i) $K:[e, f] \times[e, f] \times \mathbb{R} \rightarrow \mathbb{R}$ and $\vartheta:[e, f] \rightarrow \mathbb{R}$ are continuous functions;
(ii) there exists $\gamma \in[0,1)$ such that

$$
|K(u, v, h t(v))-K(u, v, h s(v))| \leq \gamma \frac{\max \left\{\begin{array}{c}
|t(v)-s(v)|,|h t(v)-h s(v)|, \\
|t(v)-h t(v)|,\left|h t(v)-h^{2} t(v)\right|, \\
|s(v)-h s(v)|,\left|h s(v)-h^{2} s(v)\right|
\end{array}\right\}}{f-e}
$$

for all $t, s \in C([e, f], \mathbb{R})$ and $u, v \in[e, f]$.
Then nonlinear integral equation (3.1) has a unique solution.
Proof. $W=C[e, f], \varrho(h, g)=|h-g|=\max _{t \in[e, f]}|h t-g t|$, for all $h, g \in W$, and $(W, \varrho)$ is a complete metric space. $h: W \rightarrow W$ be a continuous operator defined by

$$
h t(u)=\vartheta(u)+\int_{e}^{f} K(u, v, t(v)) d v .
$$

Starting at the point $t_{0} \in W$, the sequence $\left\{t_{n}\right\}$ is constructed by $t_{n}=h t_{n-1}=h^{n} t_{0}$, $n \geq 1$. From (3.1), we get

$$
t_{n+1}=h t_{n}(u)=\vartheta(u)+\int_{e}^{f} K\left(u, v, t_{n}(v)\right) d v
$$

Now, we will demonstrate that $h$ is a generalized $\theta$-convex contractive mapping. We can write

$$
\begin{aligned}
\left|h^{2} t(u)-h^{2} s(u)\right| & =\left|\int_{e}^{f} K(u, v, h t(v)) d v-\int_{e}^{f} K(u, v, h s(v)) d v\right| \\
& \leq \int_{e}^{f}|K(u, v, h t(v))-K(u, v, h s(v))| d v \\
& \leq \frac{\gamma}{f-e} \int_{e}^{f} \max \left\{\begin{array}{c}
|t(v)-s(v)|,|h t(v)-h s(v)| \\
|t(v)-h t(v)|,\left|h t(v)-h^{2} t(v)\right| \\
|s(v)-h s(v)|,\left|h s(v)-h^{2} s(v)\right|
\end{array}\right\} d v
\end{aligned}
$$

and

$$
\begin{aligned}
\varrho\left(h^{2} t, h^{2} s\right) & =\max _{u \in[e, f]}\left|h^{2} t(u)-h^{2} s(u)\right| \\
& \leq \frac{\gamma}{f-e} \max _{u \in[e, f]} \int_{e}^{f} \max \left\{\begin{array}{r}
|t(v)-s(v)|,|h t(v)-h s(v)| \\
|t(v)-h t(v)|,\left|h t(v)-h^{2} t(v)\right| \\
|s(v)-h s(v)|,\left|h s(v)-h^{2} s(v)\right|
\end{array}\right\} d v \\
& \leq \frac{\gamma}{f-e} \max \left[\max _{c \in[e, f]}\left\{\begin{array}{c}
|t(c)-s(c)|,|h t(c)-h s(c)|, \\
|t(c)-h t(c)|,\left|h t(c)-h^{2} t(c)\right|, \\
|s(c)-h s(c)|,\left|h s(c)-h^{2} s(c)\right|
\end{array}\right\}\right] \int_{e}^{f} d v \\
& \leq \gamma \max \left\{\varrho(t, s), \varrho(h t, h s), \varrho(t, h t), \varrho(s, h s), \varrho\left(h t, h^{2} t\right), \varrho\left(h s, h^{2} s\right)\right\} \\
& \leq \gamma M_{I}(t, s) .
\end{aligned}
$$

Thus

$$
\varrho\left(h^{2} t, h^{2} s\right) \leq \gamma M_{I}(t, s)
$$

Define $\theta(t)=e^{\sqrt{t}}$ for $t>0$ and $\varsigma(t)=\gamma^{\frac{1}{2}} t$. We have

$$
e^{\sqrt{\rho\left(h^{2} t, h^{2} s\right)}} \leq e^{\gamma^{\frac{1}{4}} \sqrt{M_{I}(t, s)}}=\left[e^{\sqrt{\varrho\left(M_{I}(t, s)\right)}}\right]^{\kappa}
$$

where $\kappa=\gamma^{\frac{1}{4}}$. Thus, we get

$$
\theta\left(\varrho\left(h^{2} t, h^{2} s\right) \leq\left[\theta\left(\varsigma\left(M_{I}(t, s)\right)\right)\right]^{\kappa}+L N_{I}(t, s)\right.
$$

where $L \geq 0$. This shows that, $h$ is a generalized $\theta$-convex contractive mapping. That is, the conditions of Theorem 1 are hold. Thus, $h$ has a unique fixed point in $W$, and so, the nonlinear integral equation (3.1) has a unique solution.

## 4. Conclusion

We present generalized $\theta$-convex contractive mappings in this paper. This contractive condition not only extends several existing contraction definitions but also merge some existing contractions. Afterward, we investigate the existence of a fixed point for our novel type contraction, we state some consequences. Our results generalize and merge the results derived by Istratescu 9, 10 and Jleli 5, and some
other connected results in the literature. Our new contraction can be considered as an interesting generalization of Darbo's fixed point problem 26, 27. As well as the corollaries in this paper, to underline the novelty of our given results, we show an example that shows that Theorem 1 is a genuine generalization of Istratescu's results 9. Moreover, as a possible application, we applied our main results to study the existence of a solution for a nonlinear integral equation. The new concept allows for further studies and applications. By choosing the appropriate auxiliary function such as simulation function and others, one can get several more results. Also, one can get the analogue of our result in the set-up of cyclic mappings.

Authors Contribution Statement The authors contributed equally to this work. All authors read and approved the final copy of this paper.

Declaration of Competing Interests Authors state no conflict of interest.

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http://communications.science.ankara.edu.tr
Commun.Fac.Sci.Univ.Ank.Ser. A1 Math. Stat.
Volume 73, Number 1, Pages 25-36 (2024)
DOI:10.31801/cfsuasmas. 1186168
ISSN 1303-5991 E-ISSN 2618-6470
Research Article; Received: October 8, 2022; Accepted: September 30, 2023

# A STRONGER FORM OF LOCALLY CLOSED SET AND ITS HOMEOMORPHIC IMAGE 

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#### Abstract

Through this paper, via the operators ( $\cdot)^{\star}$ and $\Psi$, we presented notion of $\star$-Locally set in an ideal topological space $\zeta_{\mathbb{I}}$ as a new stronger form of locally closed set, and considered relations with various existing weak form of locally closed set. Preservations of direct images as well as inverse images of $(\cdot)^{\star}, \Psi, \star$-perfect and various weak forms of locally closed set including $\star$-Locally closed set are important investigating part. Besides, we pointed out that consideration of 'bijectivity' in Lemma 3.1 of 24 is sufficient, and the Lemma 3.3 of 24 is wrong. We demonstrated two modifications of the last one.


## 1. Introduction

Locally closed set and its study is not a new idea in topology. This notion was disclosed by Bourbaki 3, and after that it has been extensively studied by a good number of mathematicians (see $[7,12,20,21)$. This study has been interesting because it generalizes both open and closed sets. But the study of a locally closed set relative to an ideal (see 13) is a new idea, and this has been introduced through this paper. The authors Jeyanthi et al. 12 and the author Dontchev 6 have studied locally closed sets in terms of ideal, but these locally closed sets differ somewhat from the current one.

We now consider some preliminary concepts from literature for developing the paper.

Consider a topological space $(\mathbf{Z}, \mathbb{T})$ (henceforth, in this paper we shall denote it by $\zeta$ ), and suppose $\mathbb{I}$ is an ideal on $\mathbf{Z}$. The set-valued map $(\cdot)^{\star}: \wp(\mathbf{Z}) \rightarrow \wp(\mathbf{Z})$ associated by the formula ' $H^{\star}=\left\{a \in \mathbf{Z}: G_{a} \cap H \notin \mathbb{I}\right.$ for every $\left.G_{a} \in \mathbb{T}_{a}\right\}$ for every $H \subseteq \mathbf{Z}^{\prime}$ is designated as the local function 11 w.r.t. the ideal $\mathbb{I}$ and the topology $\mathbb{T}$,

[^2]where $\mathbb{T}_{a}=\{G \in \mathbb{T}: a \in G\}$ and $\wp(\mathbf{Z})$ stands for power set of $\mathbf{Z}$. Other notations used instead of $H^{\star}$ are $H^{\star}(\mathbb{I}, \mathbb{T})$ and $H^{\star}(\mathbb{I})$. For the trivial ideals $\{\varnothing\}$ and $\wp(\mathbf{Z})$, values of $(\cdot)^{\star}$ are $H^{\star}(\{\varnothing\})=\mathrm{Cl}(H)$ (closure operator) and $H^{\star}(\wp(\mathbf{Z}))=\varnothing$ (zero operator), respectively. An interesting ideal on $\mathbf{Z}$ is $\mathbb{I}_{n}$ consisting of all nowhere dense sets of $\zeta$, and $H^{\star}\left(\mathbb{I}_{n}\right)=\mathrm{Cl}(\operatorname{Int}(\mathrm{Cl}(H))$ ) (see 11), where 'Int' stands for interior operator. Further, for the ideals $\mathbb{I}_{f}=\{I \subseteq \mathbf{Z}: I$ is finite $\}$ and $\mathbb{I}_{c}=\{I \subseteq$ $\mathbf{Z}: I$ is countable $\}, H^{\star}\left(\mathbb{I}_{f}\right)=H^{\omega}$ (collection of all $\omega$-accumulation point of $H$ ) and $H^{\star}\left(\mathbb{I}_{c}\right)=H^{c d}$ (collection of all condensation point of $H$ ) (see 11). Thus one can think the local function $(\cdot)^{\star}$ as a generalization of closure operator.

An important set-operator familiar to researchers as a complement of the local function $(\cdot)^{\star}: \wp(\mathbf{Z}) \rightarrow \wp(\mathbf{Z})$ is $\Psi: \wp(\mathbf{Z}) \rightarrow \wp(\mathbf{Z})$, and its value acting on $H \subseteq \mathbf{Z}$ is calculated by the formula $\Psi(H)=\mathbf{Z} \backslash(\mathbf{Z} \backslash H)^{\star} 22$. Note that $(\cdot)^{\star}$ (resp., $\Psi)$ is not necessarily a closure (resp., interior) operator. However, the operator $\mathrm{Cl}^{\star}: \wp(\mathbf{Z}) \rightarrow \wp(\mathbf{Z})$ given by the formula $\mathrm{Cl}^{\star}(H)=H \cup H^{\star}$ determines Kuratowski's closure operator $2,11,13,27$, and henceforth $\mathbf{Z}$ gets a new topology, named $\star$ topology $1,2,8-10,16,23$, induced by $\mathrm{Cl}^{\star}$. Let's name this topology as $\mathbb{T}^{\star}$. Clearly, $\mathbb{T} \subseteq \mathbb{T}^{\star}($ see 11$)$. The interior operator of the space $\zeta^{\star}=\left(\mathbf{Z}, \mathbb{T}^{\star}\right)$ is given by $\operatorname{Int}{ }^{\star}(H)=\mathbf{Z} \backslash \mathrm{Cl}^{\star}(\mathbf{Z} \backslash H)$.

Moreover, if $H \subseteq H^{\star}$, then $H$ is known as $\star$-dense in itself 10 , and if $H=H^{\star}$, then $H$ is termed as $\star$-perfect 10 .

## 2. $L^{\star}$ Operator

We are beginning this section with an example to draw interest to the fact that through idealizing a space $\zeta$ by way of a proper ideal $\mathbb{I}$ (i.e., $\mathbf{Z} \notin \mathbb{I}$ ), one can find an $H \subseteq \mathbf{Z}$ for which $H^{\star}$ intersects $\Psi(H)$ i.e., the assertion ' $K^{\star} \cap \Psi(K)=\varnothing$ for every $K \subseteq \mathbf{Z}$ ' need no longer be correct. The notations $\zeta_{\mathbb{I}}$ and $\zeta_{\mathbb{I}}^{\star}$ will be used to recognize respectively the triplets $(\mathbf{Z}, \mathbb{T}, \mathbb{I})$ and $\left(\mathbf{Z}, \mathbb{T}^{\star}, \mathbb{I}\right)$, ideal topological spaces, in this write-up.
Example 1. Consider $\mathbb{T}=\left\{\varnothing,\left\{\ell_{1}\right\}, \mathbf{Z}\right\}$ and $\mathbb{I}=\left\{\varnothing,\left\{\ell_{2}\right\}\right\}$ on $\mathbf{Z}=\left\{\ell_{1}, \ell_{2}, \ell_{3}\right\}$. Then for $H=\left\{\ell_{1}, \ell_{2}\right\}, H^{\star}=\mathbf{Z}, \Psi(H)=\left\{\ell_{1}\right\}$ and $H^{\star} \cap \Psi(H) \neq \varnothing$.

Definition 1. We define the $L^{\star}$ operator on $\zeta_{\mathbb{I}}$ as a set-valued map $L^{\star}: \wp(\mathbf{Z}) \rightarrow$ $\wp(\mathbf{Z})$ by the equation $L^{\star}(H)=H^{\star} \cap \Psi(H)$ for every $H \subseteq \mathbf{Z}$.

Remark 1. As, we know from 11 that $H^{\star}(\mathbb{I}, \mathbb{T})=H^{\star}\left(\mathbb{I}, \mathbb{T}^{\star}\right)$, so $L^{\star}$ values of every $H \subseteq \mathbf{Z}$ w.r.t. $\zeta_{\mathbb{I}}$ and $\zeta_{\mathbb{I}}^{\star}$ are same.

We shall now discuss the value of $L^{\star}(H)$ for different ideals on a topological space.

- $\mathbb{I}=\{\varnothing\}$ implies $L^{\star}(H)=\mathrm{Cl}(H) \cap(\mathbf{Z} \backslash \mathrm{Cl}(\mathbf{Z} \backslash H))=\mathrm{Cl}(H) \cap \operatorname{Int}(H)=$ $\operatorname{Int}(H)$.
- $\mathbb{I}=\wp(\mathbf{Z})$ implies $L^{\star}(H)=\varnothing \cap \Psi(H)=\varnothing$.
- $\mathbb{I}=\mathbb{I}_{n}$ implies $L^{\star}(H)=\mathrm{Cl}(\operatorname{Int}(\mathrm{Cl}(H))) \cap \operatorname{Int}(\mathrm{Cl}(\operatorname{Int}(H)))=\operatorname{Int}(\mathrm{Cl}(\operatorname{Int}(H)))$.
- $\mathbb{I}=\mathbb{I}_{f}$ implies $L^{\star}(H)=H^{\omega} \cap\left(\mathbf{Z} \backslash(\mathbf{Z} \backslash H)^{\omega}\right) \subseteq H^{\omega}$.
- $\mathbb{I}=\mathbb{I}_{c}$ implies $L^{\star}(H)=H^{c d} \cap\left(\mathbf{Z} \backslash(\mathbf{Z} \backslash H)^{c d}\right) \subseteq H^{c d}$.

Study of the $L^{\star}$ operator will be therefore fascinating if we are deal with a non-trivial ideal (non-trivial means other than $\{\varnothing\}$ and $\wp(\mathbf{Z})$ ).

Theorem 1. For $H, K \subseteq \mathbf{Z}$, the followings are true in $\zeta_{\mathbb{I}}$ :
(1) $L^{\star}(\varnothing)=\varnothing$,
(2) $L^{\star}(\mathbf{Z})=\mathbf{Z}^{\star}$,
(3) $L^{\star}(\mathbf{Z})=\mathbf{Z}$ if and only if $\mathbb{I} \cap \mathbb{T}=\{\varnothing\}$,
(4) $L^{\star}(H)=\Psi(H) \backslash \Psi(\mathbf{Z} \backslash H)$,
(5) $L^{\star}(H)=H^{\star} \backslash(\mathbf{Z} \backslash H)^{\star}$,
(6) $\mathbf{Z} \backslash L^{\star}(H)=(\mathbf{Z} \backslash H)^{\star} \cup\left(\mathbf{Z} \backslash H^{\star}\right)$,
(7) $L^{\star}(\mathbf{Z} \backslash H)=\mathbf{Z} \backslash\left(\Psi(H) \cup H^{\star}\right)$,
(8) For $H \subseteq K, L^{\star}(H) \subseteq L^{\star}(K)$,
(9) $L^{\star}(H) \cup L^{\star}(K) \subseteq L^{\star}(H \cup K)$,
(10) $L^{\star}(H \cap K) \subseteq L^{\star}(H) \cap L^{\star}(K)$,
(11) $L^{\star}(H) \subseteq H^{\star}$,
(12) $L^{\star}(H) \subseteq \Psi(H)$,
(13) $H \cap L^{\star}(H)=H^{\star} \cap \operatorname{Int}^{\star}(H)$,
(14) $H \cap L^{\star}(H) \subseteq \operatorname{Int}^{\star}(H)$,
(15) $L^{\star}(H) \subseteq H^{\star} \subseteq \mathrm{Cl}^{\star}(H) \subseteq \mathrm{Cl}(H)$,
(16) For $H \in \mathbb{T}^{\star}, H \cap H^{\star} \subseteq L^{\star}(H) \subseteq H^{\star}$,
(17) For $H \in \mathbb{T}, H \cap H^{\star} \subseteq L^{\star}(H) \subseteq H^{\star}$,
(18) For a regular open $H$ [25], $L^{\star}(H)=H \cap H^{\star}$,
(19) $\operatorname{Int}\left(L^{\star}(H)\right)=\Psi(H) \cap \operatorname{Int}\left(H^{\star}\right)$,
(20) $\operatorname{Int}^{\star}\left(L^{\star}(H)\right) \supseteq \Psi(H) \cap \operatorname{Int}^{\star}\left(H^{\star}\right)$,
(21) $\mathrm{Cl}\left(L^{\star}(H)\right) \subseteq \mathrm{Cl}(\Psi(H)) \cap H^{\star}$,
(22) $\mathrm{Cl}^{\star}\left(L^{\star}(H)\right) \subseteq \mathrm{Cl}^{\star}(\Psi(H)) \cap H^{\star}$,
(23) $\operatorname{Int}^{\star}\left(H^{\star}\right) \cap \Psi(H) \subseteq \operatorname{Int}^{\star}\left(L^{\star}(H)\right) \subseteq \mathrm{Cl}^{\star}\left(L^{\star}(H)\right) \subseteq H^{\star} \cap \mathrm{Cl}^{\star}(\Psi(H))$,
(24) For $a \star$-perfect set $H, L^{\star}(H)=H \cap \Psi(H)=\operatorname{Int}^{\star}(H)$,
(25) For $a \star$-dense in itself set $H, L^{\star}(H) \supseteq \operatorname{Int}^{\star}(H)$.

Proof. (1) $L^{\star}(\varnothing)=\varnothing^{\star} \cap \Psi(\varnothing)=\varnothing$.
(2) $L^{\star}(\mathbf{Z})=\mathbf{Z}^{\star} \cap \Psi(\mathbf{Z})=\mathbf{Z}^{\star} \cap \mathbf{Z}=\mathbf{Z}^{\star}$.
(3) Follows from the fact $\mathbf{Z}^{\star}=\mathbf{Z}$ if and only if $\mathbb{I} \cap \mathbb{T}=\{\varnothing\}$.
(4) $L^{\star}(H)=H^{\star} \cap \Psi(H)=(\mathbf{Z} \backslash \Psi(\mathbf{Z} \backslash H)) \cap \Psi(H)=\Psi(H) \backslash \Psi(\mathbf{Z} \backslash H)$.
(5) $L^{\star}(H)=H^{\star} \cap \Psi(H)=H^{\star} \cap\left(\mathbf{Z} \backslash(\mathbf{Z} \backslash H)^{\star}\right)=H^{\star} \backslash(\mathbf{Z} \backslash H)^{\star}$.
(6) $\mathbf{Z} \backslash L^{\star}(H)=\mathbf{Z} \backslash\left(H^{\star} \cap \Psi(H)\right)=\left(\mathbf{Z} \backslash H^{\star}\right) \cup(\mathbf{Z} \backslash \Psi(H))=\left(\mathbf{Z} \backslash H^{\star}\right) \cup(\mathbf{Z} \backslash H)^{\star}$.
(7) Obvious.
(8) Obvious.
(9) Follows from 8.
(10) Follows from 8.
(11) Obvious.
(12) Obvious.
(13) $H \cap L^{\star}(H)=H \cap\left(H^{\star} \cap \Psi(H)\right)=H^{\star} \cap \operatorname{Int}^{\star}(H)$.
(14) From 10, $L^{\star}(H) \subseteq \Psi(H)$. Therefore, $H \cap L^{\star}(H) \subseteq H \cap \Psi(H)=\operatorname{Int}^{\star}(H)$.
(15) Obvious from the fact $H^{\star} \subseteq H^{\star} \cup H=\mathrm{Cl}^{\star}(H) \subseteq \mathrm{Cl}(H)$.
(16) $H \in \mathbb{T}$ implies $H \subseteq \Psi(H)$. Now $L^{\star}(H)=H^{\star} \cap \Psi(H)$ implies $H^{\star} \cap H \subseteq$ $L^{\star}(H)$.
(17) Obvious from the fact $\mathbb{T} \subseteq \mathbb{T}^{\star}$.
(18) Since $H$ is regular open, so $H=\Psi(H) 2818$. Now, $L^{\star}(H)=H^{\star} \cap \Psi(H)=$ $H^{\star} \cap H$.
(19) $\operatorname{Int}\left(L^{\star}(H)\right)=\operatorname{Int}\left(\Psi(H) \cap H^{\star}\right)=\operatorname{Int}(\Psi(H)) \cap \operatorname{Int}\left(H^{\star}\right)=\Psi(H) \cap \operatorname{Int}\left(H^{\star}\right)$.
(20) $\operatorname{Int}^{\star}\left(L^{\star}(H)\right)=\operatorname{Int}^{\star}\left(H^{\star} \cap \Psi(H)\right)=\left[H^{\star} \cap \Psi(H)\right] \cap \Psi\left[H^{\star} \cap \Psi(H)\right]=\left[H^{\star} \cap\right.$ $\Psi(H)] \cap\left[\Psi\left(H^{\star}\right) \cap \Psi(\Psi(H))\right] \supseteq\left[H^{\star} \cap \Psi(H)\right] \cap\left[\Psi\left(H^{\star}\right) \cap \Psi(H)\right]=\left[H^{\star} \cap\right.$ $\left.\Psi\left(H^{\star}\right)\right] \cap \Psi(H)=\operatorname{Int}^{\star}\left(H^{\star}\right) \cap \Psi(H)$.
(21) Similar to 19.
(22) Similar to 19.
(23) Follows from 20.
(24) Trivial.
(25) Trivial.

Inequality of the result (9) of Theorem 1 is highlighted in next example.
Example 2. Take $\mathbf{Z}=\mathbb{R}$ (set of reals) with usual topology and $\mathbb{I}=\{\varnothing\}$. Pick $H=[0,2021)$ and $K=[2021,2022)$. Then $L^{\star}(H)=\operatorname{Int}(H)=(0,2021), L^{\star}(K)=$ $\operatorname{Int}(K)=(2021,2022)$ and $L^{\star}(H \cup K)=L^{\star}([0,2022))=(0,2022)$. Evidently, $L^{\star}(H) \cup L^{\star}(K) \neq L^{\star}(H \cup K)$.

Theorem 2. Suppose $\mathbb{I}$ is an ideal on $\zeta$ and $H \subseteq \mathbf{Z}$. If $a \in L^{\star}(H)$, then there exists at least one $K_{a} \in \mathbb{T}_{a}$ such that $K_{a} \notin \mathbb{I}$ but $K_{a} \backslash H \in \mathbb{I}$.

Proof. $a \in L^{\star}(H)$ gives $a \in H^{\star}$ but $a \notin(\mathbf{Z} \backslash H)^{\star}$. Now, $a \notin(\mathbf{Z} \backslash H)^{\star}$ assures the existence of a $K_{a} \in \mathbb{T}_{a}$ such that $K_{a} \cap(\mathbf{Z} \backslash H)=K_{a} \backslash H \in \mathbb{I}$. On the other hand, $a \in H^{\star}$ tells that $K_{a} \cap H \notin \mathbb{I}$. This directs that $K_{a} \notin \mathbb{I}$, since $\mathbb{I}$ is an ideal. Hence, $K_{a} \notin \mathbb{I}$ but $K_{a} \backslash H \in \mathbb{I}$, as aimed.

We talk about the validation of the converse part of Theorem 2 in next example.
Example 3. Take $\mathbb{T}=\left\{\varnothing,\left\{\ell_{1}\right\},\left\{\ell_{2}\right\}, \mathbf{Z}\right\}$ and $\mathbb{I}=\left\{\varnothing,\left\{\ell_{1}\right\}\right\}$ on $\mathbf{Z}=\left\{\ell_{1}, \ell_{2}\right\}$. Let $H=\left\{\ell_{2}\right\}$. Then $H^{\star}=\left\{\ell_{2}\right\}$ and $\Psi(H)=\mathbf{Z}$ and hence $L^{\star}(H)=\left\{\ell_{2}\right\}$. Now, pick up the point $\ell_{1}$ and choose $K_{\ell_{1}}=\mathbf{Z} \in \mathbb{T}_{\ell_{1}}$. Evidently, $K_{\ell_{1}} \notin \mathbb{I}, K_{\ell_{1}} \backslash H=\left\{\ell_{1}\right\} \in \mathbb{I}$ but $\ell_{1} \notin L^{\star}(H)$. Therefore, the reverse direction of Theorem D will usually not work.

## 3. *-Locally Closed Sets

Definition 2. We call an $H \subseteq \mathbf{Z}$ as $\star$-Locally closed in $\zeta_{\mathbb{I}}$ if there is a $K \subseteq \mathbf{Z}$ such that $H=L^{\star}(K)$, and use the symbol $L^{\star}\left(\zeta_{\mathbb{I}}\right)$ to mean $\{H \subseteq \mathbf{Z}: H$ is $\star$-Locally closed $\}$.

Example 4. Topologize $\mathbf{Z}=\mathbb{R}$ by considering $\mathbb{T}=\{\varnothing, \mathbb{Q}, \mathbb{R}\}$ and $\mathbb{I}=\wp(\mathbb{Q})$, where $\mathbb{Q}$ is the set of all rationals. Then for any $H \subseteq \mathbf{Z}$,

$$
H^{\star}= \begin{cases}\varnothing, & \text { if } H \cap(\mathbb{R} \backslash \mathbb{Q})=\varnothing \\ \mathbb{R} \backslash \mathbb{Q}, & \text { if } H \cap(\mathbb{R} \backslash \mathbb{Q}) \neq \varnothing\end{cases}
$$

Take $L=\mathbb{R} \backslash \mathbb{Q}$. We observe that $L=L^{\star} \cap \Psi(L)$. So, $\mathbb{R} \backslash \mathbb{Q}$ is $a \star$-Locally closed set.

Example 5. Consider $\zeta_{\mathbb{I}}$ discussed in Example 1, and take $H=\left\{\ell_{1}\right\}, K=\left\{\ell_{1}, \ell_{2}\right\}$. Since $H=K^{\star} \cap \Psi(K)$, so $H$ is $\star$-Locally closed in $\zeta_{\mathbb{I}}$.
Definition 3. An $L \subseteq \mathbf{Z}$ of a space $\zeta$ is familiar with the name locally closed [7] (resp., semi-locally closed [26], $\lambda$-locally closed [20]) if we can give the form $\bar{L}=H \cap K$, where $H$ is open (resp., semi-open 14, $\lambda$-open [20]) and $K$ is closed (resp., semi-closed, closed).
Definition 4. An $L \subseteq \mathbf{Z}$ is addressed as $\mathbb{I}$-locally closed [6] (resp., semi-I्I-locally closed [12]) if we can present $L$ as $L=H \cap K$, where $H \in \mathbb{T}$ and $K$ is $\star$-perfect (resp., $L=H \cap L^{\star}$, where $H$ is semi-open). An equivalent definition of $L$ to be $\mathbb{I}$-locally closed is $L=H \cap L^{\star}$, where $H \in \mathbb{T}$ (see 12]).

Remark 2. As we know from [11], $H^{\star}$ is closed, and from [22], $\Psi(H)$ is open, it is derived that $\star$-Locally closed sets are locally closed. For reverse direction, we consider next example.
Example 6. Take $\mathbb{T}=\left\{\varnothing,\left\{\ell_{1}\right\},\left\{\ell_{2}\right\},\left\{\ell_{4}\right\},\left\{\ell_{1}, \ell_{2}\right\},\left\{\ell_{1}, \ell_{4}\right\},\left\{\ell_{2}, \ell_{4}\right\},\left\{\ell_{1}, \ell_{2}, \ell_{4}\right\}, \mathbf{Z}\right\}$ and $\mathbb{I}=\left\{\varnothing,\left\{\ell_{1}\right\},\left\{\ell_{3}\right\},\left\{\ell_{1}, \ell_{3}\right\}\right\}$ on $\mathbf{Z}=\left\{\ell_{1}, \ell_{2}, \ell_{3}, \ell_{4}\right\}$. Different values of $K \subseteq \mathbf{Z}$ under the operators Cl , Int, $(\cdot)^{\star}$ and $\Psi$ are considered in TABLE 1 .

TABLE 1. Values of $K \subseteq \mathbf{Z}$ under various operators

| $K$ | $\mathrm{Cl}(K)$ | $\operatorname{Int}(K)$ | $\mathrm{Cl}(\operatorname{Int}(K))$ | $K^{\star}$ | $\Psi(K)$ | $L^{\star}(K)$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\varnothing$ | $\varnothing$ | $\varnothing$ | $\varnothing$ | $\varnothing$ | $\left\{\ell_{1}\right\}$ | $\varnothing$ |
| $\left\{\ell_{1}\right\}$ | $\left\{\ell_{1}, \ell_{3}\right\}$ | $\left\{\ell_{1}\right\}$ | $\left\{\ell_{1}, \ell_{3}\right\}$ | $\varnothing$ | $\left\{\ell_{1}\right\}$ | $\varnothing$ |
| $\left\{\ell_{2}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{2}\right\}$ |
| $\left\{\ell_{3}\right\}$ | $\left\{\ell_{3}\right\}$ | $\varnothing$ | $\varnothing$ | $\varnothing$ | $\left\{\ell_{1}\right\}$ | $\varnothing$ |
| $\left\{\ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{4}\right\}$ | $\left\{\ell_{4}\right\}$ |
| $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{1}, \ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{1}, \ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{2}\right\}$ |
| $\left\{\ell_{1}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{3}\right\}$ | $\left\{\ell_{1}\right\}$ | $\left\{\ell_{1}, \ell_{3}\right\}$ | $\varnothing$ | $\left\{\ell_{1}\right\}$ | $\varnothing$ |
| $\left\{\ell_{1}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{4}\right\}$ | $\left\{\ell_{4}\right\}$ |
| $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{2}\right\}$ |
| $\left\{\ell_{2}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\mathbf{Z}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ |
| $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{4}\right\}$ | $\left\{\ell_{4}\right\}$ |
| $\left\{\ell_{1}, \ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{1}, \ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{2}, \ell_{3}\right\}$ | $\left\{\ell_{1}, \ell_{2}\right\}$ | $\left\{\ell_{2}\right\}$ |
| $\left\{\ell_{1}, \ell_{2}, \ell_{4}\right\}$ | $\mathbf{Z}$ | $\left\{\ell_{1}, \ell_{2}, \ell_{4}\right\}$ | $\mathbf{Z}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\mathbf{Z}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ |
| $\left\{\ell_{1}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{1}, \ell_{4}\right\}$ | $\ell_{4}$ |
| $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\mathbf{Z}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ |
| $\mathbf{Z}$ | $\mathbf{Z}$ | $\mathbf{Z}$ | $\mathbf{Z}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ | $\mathbf{Z}$ | $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ |
|  |  |  |  |  |  |  |
|  |  |  |  |  |  |  |

We observe that $\left\{\ell_{3}\right\}$ is locally closed but not $\star$-Locally closed. Also, $\left\{\ell_{2}\right\}$ is $\star$ Locally closed but not $\star$-perfect whereas $\left\{\ell_{2}, \ell_{3}\right\}$ is $\star$-perfect but not $\star$-Locally closed. Further, $\left\{\ell_{3}, \ell_{4}\right\}$ is $\mathbb{I}$-locally closed but not $\star$-Locally closed; $\left\{\ell_{2}, \ell_{4}\right\}$ is semi-II-locally closed but not $\star$-Locally closed. Here, $\star$-Locally closed sets are precisely $\varnothing,\left\{\ell_{2}\right\}$, $\left\{\ell_{4}\right\}$ and $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$, and these are also $\mathbb{I}$-locally closed and hence, they are semi-$\star$-locally closed (as we know from 12 that $\mathbb{I}$-locally closed implies semi-I-locally closed). Because $\left\{\ell_{4}\right\}$ is $\star$-Locally closed is locally closed and hence, $\lambda$-locally closed (since locally closed implies $\lambda$-locally closed [20]), whereas $\{d\}$ in Example 2.3 of [20] $\lambda$-locally closed but not $\star$-Locally closed.

Following diagram will provide a transparent idea regarding different local versions of sets just discussed above:


Figure 1. Implication Diagram

Theorem 3. If $H$ be $\star$-dense in itself and $\star$-Locally closed in $\zeta_{\mathbb{I}}$, then $H$ is $\mathbb{I}$-locally closed.

Proof. Straightforward.
Corollary 1. If $H$ be $\star$-dense in itself and $\star$-Locally closed in $\zeta_{\mathbb{I}}$, then $H$ is semi-$\mathbb{I}$-locally closed.

Theorem 4. An $L \subseteq \mathbf{Z}$ is $\star$-Locally closed in $\zeta_{\mathbb{I}}$ if and only if $L=H^{\star} \backslash(\mathbf{Z} \backslash H)^{\star}$ for some $H \subseteq \mathbf{Z}$.

Proof. Immediate from Theorem 1(5).
Theorem 5. An $L \subseteq \mathbf{Z}$ is $\star$-Locally closed in $\zeta_{\mathbb{I}}$ if and only if $L=\Psi(H) \backslash \Psi(\mathbf{Z} \backslash H)$ for some $H \subseteq \mathbf{Z}$.

Proof. Immediate from Theorem 1(4).
Theorem 6. An $L \subseteq \mathbf{Z}$ is $\star$-Locally closed in $\zeta_{\mathbb{I}}$ if and only if $\mathbf{Z} \backslash L=(\mathbf{Z} \backslash H)^{\star} \cup$ $\left(\mathbf{Z} \backslash H^{\star}\right)$ for some $H \subseteq \mathbf{Z}$.

Proof. Obvious from Theorem 1(6).

It is known that in $\zeta$, open as well as closed sets are locally closed whereas in $\zeta_{\mathbb{I}}$, this occurrence need not longer be true in case of $\star$-Locally closedness. For this purpose, consider the next example.
Example 7. Think about Example 3, and pick $\left\{\ell_{1}\right\}$, a clopen set. Since no $H \subseteq \mathbf{Z}$ satisfies $\left\{\ell_{1}\right\}=H^{\star} \cap \Psi(H),\left\{\ell_{1}\right\}$ is not $\star$-Locally closed in $\zeta_{\mathbb{I}}$.
Theorem 7. If $\mathbb{I} \cap \mathbb{T}=\{\varnothing\}$, then every regular open set $i s \star$-Locally closed in $\zeta_{\mathbb{I}}$.
Proof. Pick a regular open set $H$. So $H=\Psi(H)$. Now, $\mathbb{I} \cap \mathbb{T}=\{\varnothing\}$ yields $H \subseteq H^{\star}$. Evidently, $H^{\star} \cap \Psi(H)=H$. This allows that $H \in L^{\star}\left(\zeta_{\mathbb{I}}\right)$.

Example 8. Following facts are observed in a $\zeta_{\mathbb{I}}$ :

- In Example 3. $\left\{\ell_{2}\right\}$ is $\star$-Locally closed but its complement $\left\{\ell_{1}\right\}$ is not.
- In Example 3. $\left\{\ell_{2}\right\}$ is $\star$-Locally closed but its super set $\left\{\ell_{2}, \ell_{3}\right\}$ is not.
- In Example 6. $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}$ is $\star$-Locally closed but its subset $\left\{\ell_{2}, \ell_{3}\right\}$ is not.
- In Example 6. for the subset $\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}, L^{\star}\left(\left\{\ell_{2}, \ell_{3}, \ell_{4}\right\}\right)$ is not open.
- In Example 6, for the subset $\left\{\ell_{4}\right\}, L^{\star}\left(\left\{\ell_{4}\right\}\right)$ is not closed.
- In Example 6, $\left\{\ell_{2}\right\}$ and $\left\{\ell_{4}\right\}$ are $\star$-closed but their union $\left\{\ell_{2}, \ell_{4}\right\}$ is not.

Remark 3. From above example, we say that the compilation $L^{\star}\left(\zeta_{\mathbb{I}}\right)$ usually does not form a topology, boolean algebra, generalized topology [15], ideal, filter [4] and grill 5. 17.

## 4. Homeomorphisms

Though this entire section, an ideal $\mathbb{I}$ is considered as proper, $\vartheta$ as $(\mathbf{W}, \mathbb{O})$ and $\vartheta_{\curlyvee(\mathbb{I})}$ as $(\mathbf{W}, \mathbb{O}, \curlyvee(\mathbb{I}))$.
Lemma 1. 24 If an ideal $\mathbb{I}$ on $\mathbf{Z}$ be proper and $\curlyvee: \mathbf{Z} \rightarrow \mathbf{W}$ bijective, then the ideal $\curlyvee(\mathbb{I})=\{\curlyvee(I): I \in \mathbb{I}\}$ is proper on $\mathbf{W}$.

Below, we now disclose that 'bijectivity' of $\curlyvee$ in Lemma 1 is sufficient to carry a (proper) ideal to a (proper) ideal.

Lemma 2. Suppose $\curlyvee: \mathbf{Z} \rightarrow \mathbf{W}$ is a map, and $\mathbb{I}$ an ideal on $\mathbf{Z}$. Then $\curlyvee(\mathbb{I})$ defined in Lemma 1 is an ideal on $\mathbf{W}$. Moreover, injectivity of $\curlyvee$ preserves 'properness' of I.

Proof. Firstly, $\varnothing \in \mathbb{I}$ (since an ideal) implies $\curlyvee(\varnothing) \in \curlyvee(\mathbb{I})$. But $\curlyvee(\varnothing)=\varnothing$. So, $\varnothing \in \curlyvee(\mathbb{I})$. Secondly, pick $E_{1}, E_{2} \in \curlyvee(\mathbb{I})$. Then, by the definition of $\curlyvee(\mathbb{I})$, choose $I_{1}, I_{2} \in \mathbb{I}$ such that $E_{1}=\curlyvee\left(I_{1}\right)$ and $E_{2}=\curlyvee\left(I_{2}\right)$. Now, $E_{1} \cup E_{2}=\curlyvee\left(I_{1}\right) \cup \curlyvee\left(I_{2}\right)=$ $\curlyvee\left(I_{1} \cup I_{2}\right)=\curlyvee\left(I_{3}\right)$, where $I_{3}=I_{1} \cup I_{2} \in \mathbb{I}$ (since $\mathbb{I}$ is ideal). This permits that $E_{1} \cup E_{2} \in \curlyvee(\mathbb{I})$. Lastly, take $F_{1} \subseteq F_{2}$ and $F_{2} \in \curlyvee(\mathbb{I})$. So, there is an $I \in \mathbb{I}$ such that $F_{2}=\curlyvee(I)$. Now, $F_{1} \subseteq \curlyvee(I)=\{\curlyvee(u): u \in I\}$ knocks us to construct an $I_{0} \subseteq \mathbf{Z}$ as: 'Pick those $u \in I$ whose images under $\curlyvee$ goes to $F_{1}$, and keep such $u$ in $I_{0}$ '. Thus, $I_{0}=\left\{u \in I: \curlyvee(u) \in F_{1}\right\}$. Clearly, $\curlyvee\left(I_{0}\right)=F_{1}$ and $I_{0} \subseteq I$. Because $\mathbb{I}$ is an ideal, $I \in \mathbb{I}$ implies $I_{0} \in \mathbb{I}$. This again implies $\curlyvee\left(I_{0}\right) \in \curlyvee(\mathbb{I})$ i.e., $F_{1} \in \curlyvee(\mathbb{I})$. Thus, we
finally present that $\curlyvee(\mathbb{I})$ is an ideal on $\mathbf{W}$.
For second part, suppose $\mathbb{I}$ is proper and $\curlyvee$ injective. Claim: $\curlyvee(\mathbb{I})$ is proper i.e., $\mathbf{W} \notin \curlyvee(\mathbb{I})$. If not, there exists $I \in \mathbb{I}$ such that $\curlyvee(I)=\mathbf{W}$. Now, $I \subseteq \mathbf{Z}$ implies $\mathbf{W}=$ $\curlyvee(I) \subseteq \curlyvee(\mathbf{Z}) \subseteq \mathbf{W}$ whence $\curlyvee(I)=\curlyvee(\mathbf{Z})$. This yields $\curlyvee^{-1}(\curlyvee(I))=\curlyvee^{-1}(\curlyvee(\mathbf{Z}))$ implies $I=\mathbf{Z}$ (since $\curlyvee$ is injective). So, $\mathbf{Z} \in \mathbb{I}$, a contradiction.

As consequences of the Lemma 1 we have following:
Theorem 8. Let $\curlyvee: \zeta_{\mathbb{I}} \rightarrow \vartheta$ is a homeomorphism. Then, for every $H \subseteq \mathbf{Z}$, we have
(1) $\curlyvee\left[H^{\star}(\mathbb{I})\right]=[\curlyvee(H)]^{\star}(\curlyvee(\mathbb{I}))$,
(2) $\curlyvee[\Psi(H)(\mathbb{I})]=\Psi[\curlyvee(H)](\curlyvee(\mathbb{I}))$.

Proof.
(1) Assume $v \notin[\curlyvee(H)]^{\star}(\curlyvee(\mathbb{I}))$. Pick an $E \in \mathbb{O}$ such that $v \in E$ and $E \cap \curlyvee(H) \in \curlyvee(\mathbb{I})$. Draw an $I \in \mathbb{I}$ such that $\curlyvee(I)=E \cap \curlyvee(H)$. Because $\curlyvee$ is injective, $\curlyvee^{-1}(E) \cap H=\curlyvee^{-1}(E) \cap \curlyvee^{-1}(\curlyvee(H))=\curlyvee^{-1}(E \cap \curlyvee(H))=$ $\curlyvee^{-1}(\curlyvee(I))=I \in \mathbb{I}$, where $\curlyvee^{-1}(E) \in \mathbb{T}_{\curlyvee^{-1}(v)}$ (by continuity of $\curlyvee$ ). This tells that $\curlyvee^{-1}(v) \notin H^{\star}(\mathbb{I})$, and we have $v \notin \curlyvee\left[H^{\star}(\mathbb{I})\right]$. So, $\curlyvee\left[H^{\star}(\mathbb{I})\right] \subseteq$ $[\curlyvee(H)]^{\star}(\curlyvee(\mathbb{I}))$. Reversely, pick $u \in \mathbf{W}$ such that $u \notin \curlyvee\left[H^{\star}(\mathbb{I})\right]$. Then, $\curlyvee^{-1}(u) \notin H^{\star}(\mathbb{I})$. There is $G \in \mathbb{T}_{\curlyvee^{-1}(u)}$ such that $G \cap H \in \mathbb{I}$. So, $\curlyvee(G) \cap$ $\curlyvee(H)=\curlyvee(G \cap H) \in \curlyvee(\mathbb{I})$, where $\curlyvee(G) \in \mathbb{O}_{u}$. This highlights that $u \notin$ $[\curlyvee(H)]^{*}(\curlyvee(\mathbb{I}))$. Therefore, $[\curlyvee(H)]^{*}(\curlyvee(\mathbb{I})) \subseteq \curlyvee\left[H^{\star}(\mathbb{I})\right]$. Hence, the result.
(2) $\curlyvee[\Psi(H)(\mathbb{I})]=\curlyvee\left[\mathbf{Z} \backslash(\mathbf{Z} \backslash H)^{\star}(\mathbb{I})\right]=\mathbf{W} \backslash \curlyvee\left[(\mathbf{Z} \backslash H)^{\star}(\mathbb{I})\right]=\mathbf{W} \backslash[\curlyvee(\mathbf{Z} \backslash$ $H)]^{\star}(\curlyvee(\mathbb{I}))$ (by first part) $=\mathbf{W} \backslash[\mathbf{W} \backslash \curlyvee(H)]^{*}(\curlyvee(\mathbb{I}))=\Psi[\curlyvee(H)](\curlyvee(\mathbb{I}))$.

Theorem 9. For a homeomorphism $\curlyvee: \zeta_{\mathbb{I}} \rightarrow \vartheta_{\curlyvee(\mathbb{I})}$, followings are well fulfilled:
(1) if $H$ be $\star$-perfect in $\zeta_{\mathbb{I}}$, then $\curlyvee(H)$ is $\star$-perfect in $\vartheta_{\curlyvee(\mathbb{I})}$,
(2) if $H$ be $\mathbb{I}$-locally closed in $\zeta_{\mathbb{I}}$, then $\curlyvee(H)$ is $\curlyvee(\mathbb{I})$-locally closed in $\vartheta_{\curlyvee(\mathbb{I})}$,
(3) if $H$ be semi-I-locally closed in $\zeta_{\mathbb{I}}$, then $\curlyvee(H)$ is semi- $\curlyvee(\mathbb{I})$-locally closed in $\vartheta_{\curlyvee(\mathbb{I})}$.

Proof. First two results are straightforward from Theorem 8 (1), and third one follows from Theorem 8 (1) and the fact that ' $E$ is semi-open implies $\gamma(E)$ is semi-open'.

For more homeomorphic image regarding $(\cdot)^{\star}$ and $\Psi$ operators interested readers can see 19.

Theorem 10. For a homeomorphism $\curlyvee: \zeta_{\mathbb{I}} \rightarrow \vartheta_{\curlyvee(\mathbb{I})}$ and for $H \subseteq \mathbf{Z}$, we have
(1) $\curlyvee\left[L^{\star}(H)(\mathbb{I})\right]=L^{\star}[\curlyvee(H)](\curlyvee(\mathbb{I}))$,
(2) $H \in L^{\star}\left(\zeta_{\mathbb{I}}\right)$ implies $\curlyvee(H) \in L^{\star}\left(\vartheta_{\curlyvee(\mathbb{I})}\right)$.

Proof. First one is derived from Theorem 8, and second one is a consequence of first part.

Lemma 3. 24] If an ideal $\mathbb{J}$ on $\mathbf{W}$ be proper and $\curlyvee: \mathbf{Z} \rightarrow \mathbf{W}$ surjective, then the ideal $\curlyvee^{-1}(\mathbb{J}):=\left\{\curlyvee^{-1}(J): J \in \mathbb{J}\right\}$ is proper on $\mathbf{Z}$.

Below, by presenting a sophisticated counterexample, we will show the Lemma 3 is wrong.

Example 9. Consider the map $\curlyvee: \mathbb{Z} \rightarrow \mathbb{N} \cup\{0\}$ as $x \mapsto|x|$. Here, $\mathbb{Z}$ and $\mathbb{N}$ denote the set of all integers and the set of all positive integers, respectively, and $|\cdot|$ is the modulus function. Note that $\curlyvee$ is surjective. Consider the subset $O$ of all odd positive integers, and take $\mathbb{J}=\wp(O)$. Then, $\mathbb{J}$ is a proper ideal on $\mathbb{N} \cup\{0\}$. Now, $\{1\} \in \mathbb{J}$ implies $\curlyvee^{-1}(\{1\})=\{-1,+1\} \in \curlyvee^{-1}(\mathbb{J})$. Though $\{-1\} \subseteq\{-1,+1\}$, $\{-1\} \notin \curlyvee^{-1}(\mathbb{J})$. Thus, $\curlyvee^{-1}(\mathbb{J})$ is not an ideal on $\mathbb{Z}$.

A modification of Lemma 3 is presented below:
Lemma 4. Let $\curlyvee: \mathbf{Z} \rightarrow \mathbf{W}$ be a map, and $\mathbb{J}$ an ideal on $\mathbf{W}$. Then

$$
\curlyvee^{\leftarrow}(\mathbb{J}):=\left\{E \subseteq \mathbf{Z}: E \subseteq \curlyvee^{-1}(J), J \in \mathbb{J}\right\}
$$

is an ideal on $\mathbf{Z}$. In addition, surjectivity of $\curlyvee$ preserves 'properness' of $\mathbb{J}$.
Proof. Firstly, $\varnothing \subseteq \gamma^{-1}(\varnothing)$, where $\varnothing \in \mathbb{J}$ (since an ideal) implies $\varnothing \in \gamma^{\leftarrow}(\mathbb{J})$. Secondly, take $E_{1} \subseteq E_{2}$ and $E_{2} \in \gamma^{\leftarrow}(\mathbb{J})$. There is a $J \in \mathbb{J}$ such that $E_{2} \subseteq$ $\curlyvee^{-1}(J)$, and so, $E_{1} \subseteq \curlyvee^{-1}(J)$ implies that $E_{1} \in \curlyvee^{\leftarrow}(\mathbb{J})$. Thirdly, consider $E_{1}$, $E_{2} \in \curlyvee^{\leftarrow}(\mathbb{J})$. Then, pick $J_{1}, J_{2} \in \mathbb{J}$ such that $E_{1} \subseteq \gamma^{-1}\left(J_{1}\right)$ and $E_{2} \subseteq \gamma^{-1}\left(J_{2}\right)$. Now, $E_{1} \cup E_{2} \subseteq \curlyvee^{-1}\left(J_{1}\right) \cup \curlyvee^{-1}\left(J_{2}\right)=\curlyvee^{-1}\left(J_{1} \cup J_{2}\right)$, where $J_{1} \cup J_{2} \in \mathbb{J}$ (since $\mathbb{J}$ is an ideal). Therefore, $E_{1} \cup E_{2} \in \gamma^{\leftarrow(\mathbb{J}) \text {. Thus, we demonstrate that } \gamma \leftarrow(\mathbb{J}) \text { is an }}$ ideal on $\mathbf{Z}$.
For second part, consider $\curlyvee$ is surjective and $\mathbb{J}$ proper. Claim: $\curlyvee \leftarrow(\mathbb{J})$ is proper. If not so, $\mathbf{Z} \in \Upsilon^{\leftarrow}(\mathbb{J})$. Choose $J \in \mathbb{J}$ such that $\mathbf{Z} \subseteq \curlyvee^{-1}(J)$. Because $\curlyvee$ is surjective, $\mathbf{W}=\curlyvee(\mathbf{Z}) \subseteq \curlyvee\left(\curlyvee^{-1}(J)\right)=J \subseteq \mathbf{W}$ implies $\mathbf{W}=J \in \mathbb{J}$, a contradiction.

We demonstrate another modification of Lemma 3 in next corollary:
Corollary 2. If $\curlyvee$ be bijective, then $\curlyvee^{-1}(\mathbb{J})$ of Lemma 3 coincides with $\curlyvee^{\leftarrow}(\mathbb{J})$, and hence, becomes an ideal.
Proof. It is transparent from the fact 'for each $J \in \mathbb{J}, \curlyvee^{-1}(J) \subseteq \gamma^{-1}(J)$ ' that $\gamma^{-1}(\mathbb{J}) \subseteq \curlyvee^{\leftarrow}(\mathbb{J})$. For backward part, let's pick an $E \in \gamma^{\leftarrow}(\mathbb{J})$. Then, $E \subseteq r^{-1}(J)$ for some $J \in \mathbb{J}$. Because $\curlyvee$ is surjective, $\curlyvee(E) \subseteq \curlyvee\left(\curlyvee^{-1}(J)\right)=J$ implies $\curlyvee(E) \in \mathbb{J}$. Because $\curlyvee$ is injective, $E=\curlyvee^{-1}(\curlyvee(E)) \in \curlyvee^{-1}(\mathbb{J})$. Thus, $\curlyvee_{\leftarrow}(\mathbb{J}) \subseteq \curlyvee^{-1}(\mathbb{J})$, as aimed.

As an application of Corollary 2, we have following important result:
Theorem 11. For a homeomorphism $\curlyvee: \zeta_{\gamma^{-1}(\mathbb{J})} \rightarrow \vartheta_{\mathbb{J}}$, and for $K \subseteq \mathbf{W}$, we have
(1) $\curlyvee^{-1}\left[K^{\star}(\mathbb{J})\right]=\left[\curlyvee^{-1}(K)\right]^{\star}\left(\curlyvee^{-1}(\mathbb{J})\right)$,
(2) $\curlyvee^{-1}[\Psi(K)(\mathbb{J})]=\Psi\left[\curlyvee^{-1}(K)\right]\left(\curlyvee^{-1}(\mathbb{J})\right)$.

Proof. (1) Assume $u \notin\left[\Upsilon^{-1}(K)\right]^{\star}\left(\Upsilon^{-1}(\mathbb{J})\right)$. Select an $E \in \mathbb{T}_{u}$ for which $E \cap$ $\curlyvee^{-1}(K) \in \curlyvee^{-1}(\mathbb{J})$. Draw a $J \in \mathbb{J}$ such that $E \cap \curlyvee^{-1}(K)=\curlyvee^{-1}(J)$. Because $\curlyvee$ is bijective, $\curlyvee(E) \cap K=\curlyvee(E) \cap \curlyvee\left(\curlyvee^{-1}(K)\right)=\curlyvee\left(E \cap \curlyvee^{-1}(K)\right)=$ $\curlyvee\left(\curlyvee^{-1}(J)\right)=J \in \mathbb{J}$, where continuity of $\curlyvee^{-1}$ implies $\curlyvee(E) \in \mathbb{O}_{\curlyvee(u)}$. This states that $\curlyvee(u) \notin K^{\star}(\mathbb{J})$, and this again implies $u \notin \gamma^{-1}\left[K^{\star}(\mathbb{J})\right]$. Therefore, $\curlyvee^{-1}\left[K^{\star}(\mathbb{J})\right] \subseteq\left[\gamma^{-1}(K)\right]^{\star}\left(\curlyvee^{-1}(\mathbb{J})\right)$. For reverse part, pick $v \notin$ $\curlyvee^{-1}\left[K^{\star}(\mathbb{J})\right]$. Then, $\curlyvee(v) \notin K^{\star}(\mathbb{J})$. Choose $F \in \mathbb{O}_{\curlyvee(v)}$ such that $F \cap K \in \mathbb{J}$. Continuity of $\curlyvee$ assures $\curlyvee^{-1}(F) \in \mathbb{T}_{v}$, and $\curlyvee^{-1}(F) \cap \curlyvee^{-1}(K)=\curlyvee^{-1}(F \cap$ $K) \in \curlyvee^{-1}(\mathbb{J})$. This indicates $v \notin\left[\Upsilon^{-1}(K)\right]^{\star}\left(\curlyvee^{-1}(\mathbb{J})\right)$, and consequently $\left[\curlyvee^{-1}(K)\right]^{\star}\left(\curlyvee^{-1}(\mathbb{J})\right) \subseteq \curlyvee^{-1}\left[K^{\star}(\mathbb{J})\right]$.
(2) $\curlyvee^{-1}[\Psi(K)(\mathbb{J})]=\curlyvee^{-1}\left[\mathbf{W} \backslash(\mathbf{W} \backslash K)^{\star}(\mathbb{J})\right]=\mathbf{Z} \backslash \curlyvee^{-1}\left[(\mathbf{W} \backslash K)^{\star}(\mathbb{J})\right]=\mathbf{Z} \backslash$ $\left[\curlyvee^{-1}(\mathbf{W} \backslash K)\right]^{\star}\left(\curlyvee^{-1}(\mathbb{J})\right)$ (by first part) $=\mathbf{Z} \backslash\left[\mathbf{Z} \backslash \curlyvee^{-1}(K)\right]^{\star}\left(\curlyvee^{-1}(\mathbb{J})\right)=$ $\Psi\left[\curlyvee^{-1}(K)\right]\left(\curlyvee^{-1}(\mathbb{J})\right)$.

Theorem 12. For a homeomorphism $\curlyvee: \zeta_{\gamma^{-1}(\mathbb{J})} \rightarrow \vartheta_{\mathbb{J}}$, followings are well fulfilled:
(1) if $K$ be $\star$-perfect in $\vartheta_{\mathbb{J}}$, then $\curlyvee^{-1}(K)$ is $\star$-perfect in $\zeta_{\gamma^{-1}(\mathbb{J})}$,
(2) if $K$ be $\mathbb{J}$-locally closed in $\vartheta_{\mathbb{J}}$, then $\curlyvee^{-1}(K)$ is $\curlyvee^{-1}(\mathbb{J})$-locally closed in $\zeta_{\gamma^{-1}(\mathbb{J})}$,
(3) if $K$ be semi- $\mathbb{J}$-locally closed in $\vartheta_{\mathbb{J}}$, then $\curlyvee^{-1}(K)$ is semi- $\curlyvee^{-1}(\mathbb{J})$-locally closed in $\zeta_{\gamma^{-1}(\mathbb{J})}$.

Proof. First two results are straightforward from Theorem 11 (1), and third one follows from Theorem 11 (1) and the fact that ' $F$ is semi-open implies $\gamma^{-1}(F)$ is semi-open'.

Theorem 13. For a homeomorphism $\curlyvee: \zeta_{\gamma^{-1}(\mathbb{J})} \rightarrow \vartheta_{\mathbb{J}}$, and for $K \subseteq \mathbf{W}$, we have
(1) $\curlyvee^{-1}\left[L^{\star}(K)(\mathbb{J})\right]=L^{\star}\left[\curlyvee^{-1}(K)\right]\left(\curlyvee^{-1}(\mathbb{J})\right)$,
(2) $K \in L^{\star}\left(\vartheta_{J}\right)$ implies $\curlyvee^{-1}(K) \in L^{\star}\left(\zeta_{\gamma^{-1}(\mathbb{J})}\right)$.

Proof. First one is derived from Theorem 11, and second one is a consequence of first part.

## 5. Conclusion

Kuratowski's local function ' $(\cdot)^{\star}$ ' is a generalized operator of the classic closure operator ' Cl ', and ' $\Psi$ ' operator is a generalized operator of the classic interior operator 'Int'. On the other side, one can think Bourbaki's locally closed sets are applications of the operators ' Cl ' and 'Int'. Replacing these classic operators by the updated generalized operator ' $(\cdot)^{\star}$ ' and ' $\Psi$ ', we derived a new version of locally closed set, and named $\star$-Locally closed. Example 6 and FIGURE 1 show that our $\star$-Locally closed version is a stronger form of locally closed set.

Author Contribution Statements Both the authors Jiarul Hoque and Shyamapada Modak have equal contribution in preparing this research article.

Declaration of Competing Interests The authors declare that there is no conflict of interest.

Acknowledgements The authors would like to thank the reviewers for their valuable suggestions and comments which helped to improve the paper.

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Volume 73, Number 1, Pages 37-63 (2024)
DOI:10.31801/cfsuasmas. 1249576
ISSN 1303-5991 E-ISSN 2618-6470
Research Article; Received: February 9, 2023; Accepted: September 19, 2023

# GENERALIZED BIVARIATE CONDITIONAL FIBONACCI AND LUCAS HYBRINOMIALS 

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#### Abstract

The Hybrid numbers are generalizations of complex, hyperbolic and dual numbers. In recent years, studies related with hybrid numbers have been increased significantly. In this paper, we introduce the generalized bivariate conditional Fibonacci and Lucas hybrinomials. Also, we present the Binet formula, generating functions, some significant identities, Catalan's identities and Cassini's identities of the generalized bivariate conditional Fibonacci and Lucas hybrinomials. Finally, we give more general results compared to the previous works.


## 1. Introduction

The Fibonacci and Lucas numbers are defined by

$$
F_{n}=\left\{\begin{array}{ll}
0 & \text { if } n=0  \tag{1}\\
1 & \text { if } n=1 \\
F_{n-1}+F_{n-2} & \text { if } n \geqslant 2
\end{array} \text { and } L_{n}= \begin{cases}2 & \text { if } n=0 \\
1 & \text { if } n=1 \\
L_{n-1}+L_{n-2} & \text { if } n \geqslant 2\end{cases}\right.
$$

respectively. For more information about the Fibonacci and Lucas numbers, we refer to book 9 . Until now, there have been interesting generalizations and applications of the Fibonacci and Lucas numbers $5-7,12,16$. For example, Falcon and Plaza found the general $k$-Fibonacci sequence $\left\{F_{k, n}\right\}_{n=0}^{\infty}$ by studying the recursive application of two geometrical transformations used in the well-known 4-triangle longest-edge (4TLE) partition 7. Furthermore, Edson and Yayenie 6 proposed the bi-periodic Fibonacci sequence. Also they gave generating function, the generalized Binet formula and some basic identities for $q_{n}$. By analogy to the studies 6 and 16, Bilgici 5 defined the bi-periodic Lucas numbers and he gave generating functions, the Binet formulas and some special identities for these sequences. Later,

[^3]Yılmaz et al. 18 presented generalized of Fibonacci and Lucas polynomials. Also they obtained some new algebraic properties on these numbers and polynomials. Yazlık et al. introduced a novel extension of the Fibonacci and Lucas $p$-numbers and demonstrated that these generalized Fibonacci and Lucas $p$-sequences can be simplified into various other number sequences 17. Ait-Amrane and Belbachir presented the bi-periodic $r$-Fibonacci sequence and its related family of companion sequences. They also explored the bi-periodic $r$-Lucas sequence of type $s$, where $s$ ranges from 1 to $r$, extending the classical Fibonacci and Lucas sequences. [1]. Belbachir and Bencherif 4 have generalized to bivariate polynomials of the Fibonacci and Lucas, properties obtained for Chebyshev polynomials. Ait-Amrane et al. presented a novel extension of hybrid polynomials, which combine elements of both Fibonacci and Lucas polynomials and studied various fundamental characteristics of these polynomials, including recurrence relations, generating functions, Binet formulas, summation formulas, and a matrix representation 2. Panwar and Singh 11 introduced a generalized bivariate Fibonacci-Like polynomials sequence. Bala and Verma 15 presented the generalized Bivariate bi-periodic Fibonacci polynomials.

For any nonzero real numbers $a, b, c$ and $d$, the generalization of bivariate biperiodic Fibonacci polynomial is defined as 15 ,

$$
B_{n}(x, y)=\left\{\begin{array}{ll}
a x B_{n-1}(x, y)+c y B_{n-2}(x, y), & \text { if } n \text { is even }  \tag{2}\\
b x B_{n-1}(x, y)+d y B_{n-2}(x, y), & \text { if } n \text { is odd }
\end{array}, n \geq 2\right.
$$

where $B_{0}(x, y)=0, B_{1}(x, y)=1$. Also, the authors obtained Catalan's identity, Cassini's identity, d'Ocagne identity and Gelin Cesaro identity along with Generating function and Binet's formula for the bivariate bi-periodic Fibonacci polynomial. The authors presented the generating function of the bivariate bi-periodic Fibonacci polynomial as:

$$
\begin{equation*}
G(t)=\frac{t+a x t^{2}-c y t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} . \tag{3}
\end{equation*}
$$

Moreover, they obtained Binet's formula for the bivariate bi-periodic Fibonacci polynomial as:

$$
\begin{equation*}
B_{n}(x, y)=\frac{(a x)^{1-\xi(n)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{n}{2}\right\rfloor}}\left(\frac{\beta_{1}^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{1}+(d-c) y\right)^{n-\left\lfloor\frac{n}{2}\right\rfloor}-\beta_{2}^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{2}+(d-c) y\right)^{n-\left\lfloor\frac{n}{2}\right\rfloor}}{\beta_{1}-\beta_{2}}\right) \tag{4}
\end{equation*}
$$

Then, Bala and Verma 3 defined the bivariate bi-periodic Lucas polynomials as follows:

For any nonzero real numbers $a_{1}$ and $a_{2}$, the generalization of bivariate biperiodic Lucas polynomial is defined as [3],

$$
l_{n}(x, y)= \begin{cases}a_{1} x l_{n-1}(x, y)+y l_{n-2}(x, y), & \text { if } n \text { is even }  \tag{5}\\ a_{2} x l_{n-1}(x, y)+y l_{n-2}(x, y), & \text { if } n \text { is odd }\end{cases}
$$

where, $l_{0}(x, y)=2, l_{1}(x, y)=a_{2} x$.

Özdemir 10 introduced the hybrid numbers as a new generalization of complex, hyperbolic and dual numbers. The set of hybrid numbers, denoted by $\mathbb{K}$, is defined as

$$
\begin{equation*}
\mathbb{K}=\left\{a+b \mathbf{i}+c \varepsilon+d \mathbf{h}: \quad a, b, c, d \in \mathbb{R}, \mathbf{i}^{2}=-1, \varepsilon^{2}=0, \mathbf{h}^{2}=1, \mathbf{i h}=-\mathbf{h} \mathbf{i}=\boldsymbol{\varepsilon}+\mathbf{i}\right\} . \tag{6}
\end{equation*}
$$

The following table presents products of $\mathbf{i}, \varepsilon$, and $\mathbf{h}$.
Table 1. Products of $\mathbf{i}, \varepsilon$, and $\mathbf{h}$

| $\times$ | 1 | $\mathbf{i}$ | $\varepsilon$ | $\mathbf{h}$ |
| :---: | :---: | :---: | :---: | :---: |
| 1 | 1 | $\mathbf{i}$ | $\varepsilon$ | $\mathbf{h}$ |
| $\mathbf{i}$ | $\mathbf{i}$ | -1 | $1-\mathbf{h}$ | $\varepsilon+\mathbf{i}$ |
| $\varepsilon$ | $\varepsilon$ | $\mathbf{h}+1$ | 0 | $-\varepsilon$ |
| $\mathbf{h}$ | $\mathbf{h}$ | $-\varepsilon-\mathbf{i}$ | $\varepsilon$ | 1 |

This table shows that the multiplication operation in the hybrid numbers is not commutative, but associative. Liana 13 presented the special kind of hybrid numbers, namely Horadam hybrid numbers. Then, Kızılates 8 obtained a new generalization of Fibonacci hybrid and Lucas hybrid numbers. He gave some algebraic properties of $q$-Fibonacci hybrid numbers and the $q$-Lucas hybrid numbers. Finally, Liana and Wloch 14 introduced the Fibonacci and Lucas hybrinomials, which can be considered as a generalization of the Fibonacci hybrid numbers and the Lucas hybrid numbers. Sevgi 12 defined the generalized Lucas hybrinomials with two variables. Also, he obtained the Binet formula, generating function and some properties for the generalized Lucas hybrinomials.

In the light of the above-cited recent works, some natural questions are that: can we define the bivariate conditional Fibonacci and Lucas Hybrinomials? Moreover, can we find the generating function, Binet formulas and some important identities for the bivariate conditional Fibonacci and Lucas Hybrinomials? In this study, we will investigate the answer to these questions.

This paper is structured in four section. First section includes preliminaries and literature review. In the second section, we define bivariate conditional Fibonacci hybrinomials and we give generating functions, Binet formulas and some important identities of these hybrinomials. In the third section, we discuss bivariate conditional Lucas polynomials and the bivariate conditional Lucas hybrinomials.

## 2. Generalized Bivariate Conditional Fibonacci Hybrinomials

In this section we give some identities of the generalized bivariate conditional Fibonacci hybrinomials. The next definition presents the bivariate conditional Fibonacci Hybrinomials.
Definition 1. For any variables $x, y$ and nonzero real numbers $a, b, c$ and $d$, we have

$$
\begin{equation*}
B H_{n}(x, y)=B_{n}(x, y)+\boldsymbol{i} B_{n+1}(x, y)+\varepsilon B_{n+2}(x, y)+\boldsymbol{h} B_{n+3}(x, y), \tag{7}
\end{equation*}
$$

where $B_{n}(x, y)$ was given in (2) and the initial conditions are $B H_{0}(x, y)=i+\varepsilon a x+$ $\boldsymbol{h}\left(a b x^{2}+d y\right)$ and $B H_{1}(x, y)=1+\boldsymbol{i} a x+\varepsilon\left(a b x^{2}+d y\right)+\boldsymbol{h}\left(a^{2} b x^{3}+a d x y+a c x y\right)$.

We can see from the following table that the generalized bivariate conditional Fibonacci hybrinomials are the generalization of many works for different values of $a, b, c$ and $d$.

Table 2. The generalized bivariate conditional Fibonacci hybrinomials

| $\boldsymbol{a}$ | $\boldsymbol{b}$ | $\boldsymbol{c}$ | $\boldsymbol{d}$ | Generalized Bivariate Conditional Fibonacci Hybrinomials |
| :---: | :---: | :---: | :---: | :---: |
| 1 | 1 | 1 | 1 | Bivariate Fibonacci Hybrinomials |
| $a$ | $b$ | 1 | 1 | Bivariate Conditional Fibonacci Hybrinomials |
| 2 | 2 | 1 | 1 | Bivariate Pell Hybrinomials |
| 1 | 1 | 2 | 2 | Bivariate Jacobsthal Hybrinomials |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |

Lemma 1. For the generalized bivariate conditional Fibonacci hybrinomials $\left\{B H_{n}(x, y)\right\}_{n=0}^{\infty}$, we have

$$
\begin{aligned}
& B H_{2 n}(x, y)=\left(a b x^{2}+(c+d) y\right) B H_{2 n-2}(x, y)-c d y^{2} B H_{2 n-4}(x, y) \\
& B H_{2 n+1}(x, y)=\left(a b x^{2}+(c+d) y\right) B H_{2 n-1}(x, y)-c d y^{2} B H_{2 n-3}(x, y)
\end{aligned}
$$

Proof. By using the definition of the generalized bivariate conditional Fibonacci hybrinomials, we obtain

$$
\begin{aligned}
B H_{2 n}(x, y)= & B_{2 n}(x, y)+\mathbf{i} B_{2 n+1}(x, y)+\varepsilon B_{2 n+2}(x, y)+\mathbf{h} B_{2 n+3}(x, y) \\
= & \left(a x B_{2 n-1}(x, y)+c y B_{2 n-2}(x, y)\right)+\mathbf{i}\left(b x B_{2 n}(x, y)+d y B_{2 n-1}(x, y)\right) \\
& +\varepsilon\left(a x B_{2 n+1}(x, y)+c y B_{2 n}(x, y)\right) \\
& +\mathbf{h}\left(b x B_{2 n+2}(x, y)+d y B_{2 n+1}(x, y)\right) \\
= & {\left[a x\left(b x B_{2 n-2}(x, y)+d y B_{2 n-3}(x, y)\right)+c y B_{2 n-2}(x, y)\right] } \\
& +\mathbf{i}\left[b x\left(a x B_{2 n-1}(x, y)+c y B_{2 n-2}(x, y)\right)+d y B_{2 n-1}(x, y)\right] \\
& \left.+\varepsilon\left[a x\left(b x B_{2 n}(x, y)+d y B_{2 n-1}(x, y)\right)+c y B_{2 n}(x, y)\right)\right] \\
& +\mathbf{h}\left[b x\left(a x B_{2 n+1}(x, y)+c y B_{2 n}(x, y)\right)+d y B_{2 n+1}(x, y)\right] \\
= & {\left[\left(a b x^{2}+c y\right) B_{2 n-2}(x, y)+d y\left(a x B_{2 n-3}(x, y)\right)\right] } \\
& +\mathbf{i}\left[\left(a b x^{2}+d y\right) B_{2 n-1}(x, y)+c y\left(b x B_{2 n-2}(x, y)\right)\right] \\
& +\varepsilon\left[\left(a b x^{2}+c y\right) B_{2 n}(x, y)+d y\left(a x B_{2 n-1}(x, y)\right)\right] \\
& +\mathbf{h}\left[\left(a b x^{2}+d y\right) B_{2 n+1}(x, y)+c y\left(b x B_{2 n}(x, y)\right)\right]
\end{aligned}
$$

$$
\begin{aligned}
= & {\left[\left(a b x^{2}+c y\right) B_{2 n-2}(x, y)+d y\left(B_{2 n-2}(x, y)-c y B_{2 n-4}(x, y)\right)\right] } \\
& +\mathbf{i}\left[\left(a b x^{2}+d y\right) B_{2 n-1}(x, y)+c y\left(B_{2 n-1}(x, y)-d y B_{2 n-3}(x, y)\right)\right] \\
& +\varepsilon\left[\left(a b x^{2}+c y\right) B_{2 n}(x, y)+d y\left(B_{2 n}(x, y)-c y B_{2 n-2}(x, y)\right)\right] \\
& +\mathbf{h}\left[\left(a b x^{2}+d y\right) B_{2 n+1}(x, y)+c y\left(B_{2 n+1}(x, y)-d y B_{2 n-1}(x, y)\right)\right] \\
= & {\left[\left(a b x^{2}+(c+d) y\right) B_{2 n-2}(x, y)-c d y^{2} B_{2 n-4}(x, y)\right] } \\
& +\mathbf{i}\left[\left(a b x^{2}+(c+d) y\right) B_{2 n-1}(x, y)-c d y^{2} B_{2 n-3}(x, y)\right] \\
& +\varepsilon\left[\left(a b x^{2}+(c+d) y\right) B_{2 n}(x, y)-c d y^{2} B_{2 n-2}(x, y)\right] \\
& +\mathbf{h}\left[\left(a b x^{2}+(c+d) y\right) B_{2 n+1}(x, y)-c d y^{2} B_{2 n-1}(x, y)\right] \\
= & \left(a b x^{2}+(c+d) y\right)\left[B_{2 n-2}(x, y)+\mathbf{i} B_{2 n-1}(x, y)+\varepsilon B_{2 n}(x, y)+\mathbf{h} B_{2 n+1}(x, y)\right] \\
& -c d y^{2}\left[B_{2 n-4}(x, y)+\mathbf{i} B_{2 n-3}(x, y)+\varepsilon B_{2 n-2}(x, y)+\mathbf{h} B_{2 n-1}(x, y)\right] \\
= & \left(a b x^{2}+(c+d) y\right) B H_{2 n-2}(x, y)-c d y^{2} B H_{2 n-4}(x, y) .
\end{aligned}
$$

Similar to the above steps, we can obtain

$$
B H_{2 n+1}(x, y)=\left(a b x^{2}+(c+d) y\right) B H_{2 n-1}(x, y)-c d y^{2} B H_{2 n-3}(x, y)
$$

Thus, the proof is completed.
Next, we give the generating function of the bivariate conditional Fibonacci hybrinomial $B H_{n}(x, y)$.

Theorem 1. The generating function for the bivariate conditional Fibonacci hybrinomial $B H_{n}(x, y)$ is

$$
\begin{align*}
\mathfrak{G}(t)=\sum_{n=0}^{\infty} B H_{n}(x, y) t^{n}= & \frac{B H_{0}(x, y)+B H_{1}(x, y) t}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \\
& +\frac{\left[B H_{2}(x, y)-\left(a b x^{2}+(c+d) y\right) B H_{0}(x, y)\right] t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}}  \tag{8}\\
& +\frac{\left[B H_{3}(x, y)-\left(a b x^{2}+(c+d) y\right) B H_{1}(x, y)\right] t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}}
\end{align*}
$$

Proof. We define

$$
\begin{aligned}
\mathfrak{G}_{0}(t) & =\sum_{n=0}^{\infty} B H_{2 n}(x, y) t^{2 n} \\
\mathfrak{G}_{1}(t) & =\sum_{n=0}^{\infty} B H_{2 n+1}(x, y) t^{2 n+1}
\end{aligned}
$$

So that

$$
\mathfrak{G}(t)=\mathfrak{G}_{0}(t)+\mathfrak{G}_{1}(t)
$$

We have

$$
\begin{aligned}
\mathfrak{G}_{0}(t)= & \sum_{n=0}^{\infty} B H_{2 n}(x, y) t^{2 n} \\
= & \sum_{n=0}^{\infty} B H_{2 n}(x, y) t^{2 n}=B H_{0}(x, y) t^{0}+B H_{2}(x, y) t^{2}+\sum_{n=2}^{\infty} B H_{2 n}(x, y) t^{2 n} \\
= & B H_{0}(x, y)+B H_{2}(x, y) t^{2} \\
& +\sum_{n=2}^{\infty}\left[\left(a b x^{2}+(c+d) y\right) B H_{2 n-2}(x, y)-c d y^{2} B H_{2 n-4}(x, y)\right] t^{2 n} \\
= & B H_{0}(x, y)+B H_{2}(x, y) t^{2}+\left(a b x^{2}+(c+d) y\right) t^{2} \sum_{n=2}^{\infty} B H_{2 n-2}(x, y) t^{2 n-2} \\
& -c d y^{2} t^{4} \sum_{n=2}^{\infty} B H_{2 n-4}(x, y) t^{2 n-4} \\
= & B H_{0}(x, y)+B H_{2}(x, y) t^{2} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2} \\
& \times\left[\sum_{n=2}^{\infty} B H_{2 n-2}(x, y) t^{2 n-2}+B H_{0}(x, y) t^{0}-B H_{0}(x, y) t^{0}\right] \\
& -c d y^{2} t^{4} \mathfrak{G}_{0}(t) \\
= & B H_{0}(x, y)+B H_{2}(x, y) t^{2}+\left(a b x^{2}+(c+d) y\right) t^{2} \mathfrak{G}_{0}(t) \\
& -\left(a b x^{2}+(c+d) y\right) t^{2} B H_{0}(x, y)-c d y^{2} t^{4} \mathfrak{G}_{0}(t) .
\end{aligned}
$$

Thus, we get

$$
\begin{equation*}
\mathfrak{G}_{0}(t)=\frac{B H_{0}(x, y)+\left(B H_{2}(x, y)-\left(a b x^{2}+(c+d) y\right) B H_{0}(x, y)\right) t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{9}
\end{equation*}
$$

Similarly, we find

$$
\begin{aligned}
\mathfrak{G}_{1}(t) & =\sum_{n=0}^{\infty} B H_{2 n+1}(x, y) t^{2 n+1} \\
& =\sum_{n=0}^{\infty} B H_{2 n+1}(x, y) t^{2 n+1} \\
& =B H_{1}(x, y) t+B H_{3}(x, y) t^{3}+\sum_{n=2}^{\infty} B H_{2 n+1}(x, y) t^{2 n+1}
\end{aligned}
$$

$$
\begin{aligned}
= & B H_{1}(x, y) t+B H_{3}(x, y) t^{3} \\
& +\sum_{n=2}^{\infty}\left[\left(a b x^{2}+(c+d) y\right) B H_{2 n-1}(x, y)-c d y^{2} B H_{2 n-3}(x, y)\right] t^{2 n+1} \\
= & B H_{1}(x, y) t+B H_{3}(x, y) t^{3}+\left(a b x^{2}+(c+d) y\right) t^{2} \sum_{n=2}^{\infty} B H_{2 n-1}(x, y) t^{2 n-1} \\
& -c d y^{2} t^{4} \sum_{n=2}^{\infty} B H_{2 n-3}(x, y) t^{2 n-3} \\
= & B H_{1}(x, y) t+B H_{3}(x, y) t^{3} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2}\left[\sum_{n=2}^{\infty} B H_{2 n-1}(x, y) t^{2 n-1}+B H_{1}(x, y) t-B H_{1}(x, y) t\right] \\
& -c d y^{2} t^{4} \mathfrak{G}_{1}(t) \\
= & B H_{1}(x, y) t+B H_{3}(x, y) t^{3}+\left(a b x^{2}+(c+d) y\right) t^{2} \mathfrak{G}_{1}(t) \\
& -\left(a b x^{2}+(c+d) y\right) t^{3} B H_{1}(x, y)-c d y^{2} t^{4} \mathfrak{G}_{1}(t) .
\end{aligned}
$$

Therefore, we get

$$
\begin{equation*}
\mathfrak{G}_{1}(t)=\frac{B H_{1}(x, y) t+\left(B H_{3}(x, y)-\left(a b x^{2}+(c+d) y\right) B H_{1}(x, y)\right) t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{10}
\end{equation*}
$$

By virtue of (9) and (10), we can obtain

$$
\begin{aligned}
\mathfrak{G}(t)= & \mathfrak{G}_{0}(t)+\mathfrak{G}_{1}(t) \\
= & \sum_{n=0}^{\infty} B H_{n}(x, y) t^{n} \\
= & \frac{B H_{0}(x, y)+B H_{1}(x, y) t+\left[B H_{2}(x, y)-\left(a b x^{2}+(c+d) y\right) B H_{0}(x, y)\right] t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \\
& \quad+\frac{\left[B H_{3}(x, y)-\left(a b x^{2}+(c+d) y\right) B H_{1}(x, y)\right] t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}}
\end{aligned}
$$

Hence, the proof is completed.
Now we give the Binet formula of the bivariate conditional Fibonacci hybrinomial $B H_{n}(x, y)$.

Theorem 2. The $n^{\text {th }}$ term of the generalized bivariate conditional Fibonacci hybrinomial $B H_{n}(x, y)$ is

$$
\begin{equation*}
B H_{n}(x, y)=\frac{\widehat{\alpha}_{\xi(n)} \beta_{1}^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{1}+(d-c) y\right)^{\left\lfloor\frac{n}{2}\right\rfloor+\xi(n)}-\widehat{\gamma}_{\xi(n)} \beta_{2}^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{2}+(d-c) y\right)^{\left\lfloor\frac{n}{2}\right\rfloor+\xi(n)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)} . \tag{11}
\end{equation*}
$$

where $\beta_{1}$ and $\beta_{2}$ roots of the characteristic equation $\lambda^{2}-\left(a b x^{2}+(c-d) y\right) \lambda-a b d x^{2} y=$ 0. Also,

$$
\begin{aligned}
\widehat{\alpha}_{\xi(n)}= & (a x)^{\xi(n+1)}+i \frac{(a x)^{\xi(n)} \beta_{1}^{\xi(n)}}{\left(a b x^{2}\right)^{\xi(n)}}\left(\beta_{1}+(d-c) y\right)^{\xi(n+1)} \\
& +\varepsilon \frac{(a x)^{\xi(n+1)} \beta_{1}}{\left(a b x^{2}\right)}\left(\beta_{1}+(d-c) y\right) \\
& +\boldsymbol{h} \frac{(a x)^{\xi(n)} \beta_{1}^{\xi(n)+1}}{\left(a b x^{2}\right)^{\xi(n)+1}}\left(\beta_{1}+(d-c) y\right)^{\xi(n+1)+1}
\end{aligned}
$$

and

$$
\begin{aligned}
\widehat{\gamma}_{\xi(n)}= & (a x)^{\xi(n+1)}+i \frac{(a x)^{\xi(n)} \beta_{2}^{\xi(n)}}{\left(a b x^{2}\right)^{\xi(n)}}\left(\beta_{2}+(d-c) y\right)^{\xi(n+1)} \\
& +\varepsilon \frac{(a x)^{\xi(n+1)} \beta_{2}}{\left(a b x^{2}\right)}\left(\beta_{2}+(d-c) y\right) \\
& +\boldsymbol{h} \frac{(a x)^{\xi(n)} \beta_{2}^{\xi(n)+1}}{\left(a b x^{2}\right)^{\xi(n)+1}}\left(\beta_{2}+(d-c) y\right)^{\xi(n+1)+1}
\end{aligned}
$$

Proof. We use the following properties throughout the proof:

- $\beta_{1}+\beta_{2}=a b x^{2}+(c-d) y$
- $\beta_{1} \cdot \beta_{2}=-a b d x^{2} y$
- $\left(\beta_{1}+d y\right)\left(\beta_{2}+d y\right)=c d y^{2}$
- $\left(\beta_{1}+d y\right)\left(a b x^{2}\right)=\beta_{1}\left(\beta_{1}+(d-c) y\right)$
- $\left(\beta_{2}+d y\right)\left(a b x^{2}\right)=\beta_{2}\left(\beta_{2}+(d-c) y\right)$.

Note that $\beta_{1}(x, y)=\beta_{1}$ and $\beta_{2}(x, y)=\beta_{2}$. By using (4), we have

$$
\begin{aligned}
B H_{2 n}(x, y)= & B_{2 n}(x, y)+\mathbf{i} B_{2 n+1}(x, y)+\varepsilon B_{2 n+2}(x, y)+\mathbf{h} B_{2 n+3}(x, y) \\
= & \frac{(a x)}{\left(a b x^{2}\right)^{n}}\left[\frac{\beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n}-\beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n}}{\beta_{1}-\beta_{2}}\right] \\
& +\mathbf{i} \frac{1}{\left(a b x^{2}\right)^{n}}\left[\frac{\beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n+1}-\beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\beta_{1}-\beta_{2}}\right] \\
& +\varepsilon \frac{(a x)}{\left(a b x^{2}\right)^{n+1}}\left[\frac{\beta_{1}^{n+1}\left(\beta_{1}+(d-c) y\right)^{n+1}-\beta_{2}^{n+1}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\beta_{1}-\beta_{2}}\right] \\
& +\mathbf{h} \frac{1}{\left(a b x^{2}\right)^{n+1}}\left[\frac{\beta_{1}^{n+1}\left(\beta_{1}+(d-c) y\right)^{n+2}-\beta_{2}^{n+1}\left(\beta_{2}+(d-c) y\right)^{n+2}}{\beta_{1}-\beta_{2}}\right]
\end{aligned}
$$

$$
\begin{aligned}
= & \frac{\beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)} \\
& \times\left[a x+\mathbf{i}\left(\beta_{1}+(d-c) y\right)+\varepsilon \frac{a x}{a b x^{2}} \beta_{1}\left(\beta_{1}+(d-c) y\right)+\mathbf{h} \frac{1}{a b x^{2}} \beta_{1}\left(\beta_{1}+(d-c) y\right)^{2}\right] \\
& -\frac{\beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)} \\
& \times\left[a x+\mathbf{i}\left(\beta_{2}+(d-c) y\right)+\varepsilon \frac{a x}{a b x^{2}} \beta_{2}\left(\beta_{2}+(d-c) y\right)+\mathbf{h} \frac{1}{a b x^{2}} \beta_{2}\left(\beta_{2}+(d-c) y\right)^{2}\right]
\end{aligned}
$$

Here, we choose the $\widehat{\alpha}_{0}$ and $\widehat{\gamma}_{0}$ as follows:

$$
\begin{aligned}
& \widehat{\alpha}_{0}=\left[a x+\mathbf{i}\left(\beta_{1}+(d-c) y\right)+\varepsilon \frac{a x}{a b x^{2}} \beta_{1}\left(\beta_{1}+(d-c) y\right)+\mathbf{h} \frac{1}{a b x^{2}} \beta_{1}\left(\beta_{1}+(d-c) y\right)^{2}\right] \\
& \widehat{\gamma}_{0}=\left[a x+\mathbf{i}\left(\beta_{2}+(d-c) y\right)+\varepsilon \frac{a x}{a b x^{2}} \beta_{2}\left(\beta_{2}+(d-c) y\right)+\mathbf{h} \frac{1}{a b x^{2}} \beta_{2}\left(\beta_{2}+(d-c) y\right)^{2}\right] .
\end{aligned}
$$

Finally, the following equation is obtained:

$$
\begin{equation*}
B H_{2 n}(x, y)=\frac{\widehat{\alpha}_{0} \beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n}-\widehat{\gamma}_{0} \beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)} \tag{12}
\end{equation*}
$$

In a similar way, by using (4), we have

$$
\begin{aligned}
B H_{2 n+1}(x, y)= & B_{2 n+1}(x, y)+\mathbf{i} B_{2 n+2}(x, y)+\varepsilon B_{2 n+3}(x, y)+\mathbf{h} B_{2 n+4}(x, y) \\
= & \frac{1}{\left(a b x^{2}\right)^{n}}\left[\frac{\beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n+1}-\beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\beta_{1}-\beta_{2}}\right] \\
& +\mathbf{i} \frac{(a x)}{\left(a b x^{2}\right)^{n+1}}\left[\frac{\beta_{1}^{n+1}\left(\beta_{1}+(d-c) y\right)^{n+1}-\beta_{2}^{n+1}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\beta_{1}-\beta_{2}}\right] \\
& +\varepsilon \frac{1}{\left(a b x^{2}\right)^{n+1}}\left[\frac{\beta_{1}^{n+1}\left(\beta_{1}+(d-c) y\right)^{n+2}-\beta_{2}^{n+1}\left(\beta_{2}+(d-c) y\right)^{n+2}}{\beta_{1}-\beta_{2}}\right] \\
& +\mathbf{h} \frac{(a x)}{\left(a b x^{2}\right)^{n+2}}\left[\frac{\beta_{1}^{n+2}\left(\beta_{1}+(d-c) y\right)^{n+2}-\beta_{2}^{n+2}\left(\beta_{2}+(d-c) y\right)^{n+2}}{\beta_{1}-\beta_{2}}\right] \\
= & \frac{\beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n+1}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)} \\
& \times\left[1+\mathbf{i} \frac{a x}{a b x^{2}} \beta_{1}+\varepsilon \frac{1}{a b x^{2}} \beta_{1}\left(\beta_{1}+(d-c) y\right)+\mathbf{h} \frac{a x}{\left(a b x^{2}\right)^{2}} \beta_{1}^{2}\left(\beta_{1}+(d-c) y\right)\right] \\
& -\frac{\beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\left(a b x^{2}\right)^{n}\left(\beta_{2}-\beta_{2}\right)} \\
& \times\left[1+\mathbf{i} \frac{a x}{a b x^{2}} \beta_{2}+\varepsilon \frac{1}{a b x^{2}} \beta_{2}\left(\beta_{2}+(d-c) y\right)+\mathbf{h} \frac{a x}{\left(a b x^{2}\right)^{2}} \beta_{2}^{2}\left(\beta_{2}+(d-c) y\right)\right] .
\end{aligned}
$$

Here, we choose the $\widehat{\alpha}_{1}$ and $\widehat{\gamma}_{1}$ as follows;

$$
\begin{aligned}
& \widehat{\alpha}_{1}=\left[1+\mathbf{i} \frac{a x}{a b x^{2}} \beta_{1}+\varepsilon \frac{1}{a b x^{2}} \beta_{1}\left(\beta_{1}+(d-c) y\right)+\mathbf{h} \frac{a x}{\left(a b x^{2}\right)^{2}} \beta_{1}^{2}\left(\beta_{1}+(d-c) y\right)\right] \\
& \widehat{\gamma}_{1}=\left[1+\mathbf{i} \frac{a x}{a b x^{2}} \beta_{2}+\varepsilon \frac{1}{a b x^{2}} \beta_{2}\left(\beta_{2}+(d-c) y\right)+\mathbf{h} \frac{a x}{\left(a b x^{2}\right)^{2}} \beta_{2}^{2}\left(\beta_{2}+(d-c) y\right)\right]
\end{aligned}
$$

Finally, the following equation is obtained.

$$
\begin{equation*}
B H_{2 n+1}(x, y)=\frac{\widehat{\alpha}_{1} \beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n+1}-\widehat{\gamma}_{1} \beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)} \tag{13}
\end{equation*}
$$

By virtue of 12 and $\sqrt{13}$, we can obtain the following equation.

$$
B H_{n}(x, y)=\frac{\widehat{\alpha}_{\xi(n)} \beta_{1}^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{1}+(d-c) y\right)^{\left\lfloor\frac{n}{2}\right\rfloor+\xi(n)}-\widehat{\gamma}_{\xi(n)} \beta_{2}^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{2}+(d-c) y\right)^{\left\lfloor\frac{n}{2}\right\rfloor+\xi(n)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{n}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)}
$$

where $\beta_{1}$ and $\beta_{2}$ roots of the characteristic equation $\lambda^{2}-\left(a b x^{2}+(c-d) y\right) \lambda-$ $a b d x^{2} y=0$. Also,

$$
\begin{aligned}
\widehat{\alpha}_{\xi(n)}= & (a x)^{\xi(n+1)}+\mathbf{i} \frac{(a x)^{\xi(n)} \beta_{1}^{\xi(n)}}{\left(a b x^{2}\right)^{\xi(n)}}\left(\beta_{1}+(d-c) y\right)^{\xi(n+1)} \\
& +\varepsilon \frac{(a x)^{\xi(n+1)} \beta_{1}}{\left(a b x^{2}\right)}\left(\beta_{1}+(d-c) y\right) \\
& +\mathbf{h} \frac{(a x)^{\xi(n)} \beta_{1}^{\xi(n)+1}}{\left(a b x^{2}\right)^{\xi(n)+1}}\left(\beta_{1}+(d-c) y\right)^{\xi(n+1)+1}
\end{aligned}
$$

and

$$
\begin{aligned}
\widehat{\gamma}_{\xi(n)}= & (a x)^{\xi(n+1)}+\mathbf{i} \frac{(a x)^{\xi(n)} \beta_{2}^{\xi(n)}}{\left(a b x^{2}\right)^{\xi(n)}}\left(\beta_{2}+(d-c) y\right)^{\xi(n+1)} \\
& +\varepsilon \frac{(a x)^{\xi(n+1)} \beta_{2}}{\left(a b x^{2}\right)}\left(\beta_{2}+(d-c) y\right) \\
& +\mathbf{h} \frac{(a x)^{\xi(n)} \beta_{2}^{\xi(n)+1}}{\left(a b x^{2}\right)^{\xi(n)+1}}\left(\beta_{2}+(d-c) y\right)^{\xi(n+1)+1}
\end{aligned}
$$

Now, we give the Catalan's identity of the bivariate conditional Fibonacci hybrinomial $B H_{n}(x, y)$.

Theorem 3. For any integers $n$ and $r$ and $n \geqslant r \geqslant 0$, we have

$$
\begin{aligned}
& B H_{2(n+r)+\xi(i)}(x, y) B H_{2(n-r)+\xi(i)}(x, y)-\left(B H_{2 n+\xi(i)}(x, y)\right)^{2} \\
& =\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\alpha}_{\xi(i)} \widehat{\gamma}_{\xi(i)} \beta_{1}{ }^{n} \beta_{2}{ }^{n}\left(\beta_{1}+(d-c) y\right)^{n+\xi(i)}\left(\beta_{2}+(d-c) y\right)^{n+\xi(i)}\left[1-\left(\frac{\beta_{1}\left(\beta_{1}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right] \\
& \quad+\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\gamma}_{\xi(i)} \widehat{\alpha}_{\xi(i)} \beta_{2}{ }^{n} \beta_{1}{ }^{n}\left(\beta_{2}+(d-c) y\right)^{n+\xi(i)}\left(\beta_{1}+(d-c) y\right)^{n+\xi(i)}\left[1-\left(\frac{\beta_{2}\left(\beta_{2}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right]
\end{aligned}
$$

where $\widehat{\alpha}_{\xi(i)}$ and $\widehat{\gamma}_{\xi(i)}$ are defined in Theorem(2) and $i \in\{0,1\}$.
Proof. In order to prove Catalan's identity, we will examine in two different cases. Case $i=0$ :

$$
\begin{align*}
B H_{2(n+r)}(x, y)= & \frac{\widehat{\alpha}_{\xi(2 n+2 r)} \beta_{1}^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor}\left(\beta_{1}+(d-c) y\right)^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor+\xi(2 n+2 r)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)} \\
& -\frac{\widehat{\gamma}_{\xi(2 n+2 r)} \beta_{2}^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor}\left(\beta_{2}+(d-c) y\right)^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor+\xi(2 n+2 r)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)}  \tag{14}\\
= & \frac{\widehat{\alpha}_{0} \beta_{1}^{n+r}\left(\beta_{1}+(d-c) y\right)^{n+r}-\widehat{\gamma}_{0} \beta_{2}^{n+r}\left(\beta_{2}+(d-c) y\right)^{n+r}}{\left(a b x^{2}\right)^{n+r}\left(\beta_{1}-\beta_{2}\right)}
\end{align*}
$$

$$
\begin{align*}
B H_{2(n-r)}(x, y)= & \frac{\widehat{\alpha}_{\xi(2 n-2 r)} \beta_{1}^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor}\left(\beta_{1}+(d-c) y\right)^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor+\xi(2 n-2 r)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)} \\
& -\frac{\widehat{\gamma}_{\xi(2 n-2 r)^{2}} \beta_{2}^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor}\left(\beta_{2}+(d-c) y\right)^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor+\xi(2 n-2 r)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)}  \tag{15}\\
= & \frac{\widehat{\alpha}_{0} \beta_{1}^{n-r}\left(\beta_{1}+(d-c) y\right)^{n-r}-\widehat{\gamma}_{0} \beta_{2}^{n-r}\left(\beta_{2}+(d-c) y\right)^{n-r}}{\left(a b x^{2}\right)^{n-r}\left(\beta_{1}-\beta_{2}\right)}
\end{align*}
$$

$$
B H_{2 n}(x, y)=\frac{\widehat{\alpha}_{\xi(2 n)} \beta_{1}^{\left\lfloor\frac{2 n}{2}\right\rfloor}\left(\beta_{1}+(d-c) y\right)^{\left\lfloor\frac{2 n}{2}\right\rfloor+\xi(2 n-2 r)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{2 n}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)}
$$

$$
\begin{equation*}
-\frac{\widehat{\gamma}_{\xi(2 n)} \beta_{2}^{\left\lfloor\frac{2 n}{2}\right\rfloor}\left(\beta_{2}+(d-c) y\right)^{\left\lfloor\frac{2 n}{2}\right\rfloor+\xi(2 n)}}{\left(a b x^{2}\right)^{\left\lfloor\frac{2 n}{2}\right\rfloor}\left(\beta_{1}-\beta_{2}\right)} \tag{16}
\end{equation*}
$$

$$
=\frac{\widehat{\alpha}_{0} \beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n}-\widehat{\gamma}_{0} \beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)}
$$

By virtue of (14), (15) and (16), we have

$$
\begin{aligned}
& B H_{2(n+r)}(x, y) B H_{2(n-r)}(x, y)-\left(B H_{2 n}(x, y)\right)^{2} \\
& =\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\alpha}_{0} \widehat{\gamma}_{0} \beta_{1}{ }^{n}{\beta_{2}}^{n}\left(\beta_{1}+(d-c) y\right)^{n}\left(\beta_{2}+(d-c) y\right)^{n}\left[1-\left(\frac{\beta_{1}\left(\beta_{1}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right] \\
& \quad+\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\gamma}_{0} \widehat{\alpha}_{0} \beta_{2}{ }^{n} \beta_{1}{ }^{n}\left(\beta_{2}+(d-c) y\right)^{n}\left(\beta_{1}+(d-c) y\right)^{n}\left[1-\left(\frac{\beta_{2}\left(\beta_{2}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right] .
\end{aligned}
$$

Case $i=1$

$$
\begin{gather*}
B H_{2(n+r)+1}(x, y)=\frac{\widehat{\alpha}_{1} \beta_{1}^{n+r}\left(\beta_{1}+(d-c) y\right)^{n+r+1}-\widehat{\gamma}_{1} \beta_{2}^{n+r}\left(\beta_{2}+(d-c) y\right)^{n+r+1}}{\left(a b x^{2}\right)^{n+r}\left(\beta_{1}-\beta_{2}\right)}  \tag{17}\\
B H_{2(n-r)+1}(x, y)=\frac{\widehat{\alpha}_{1} \beta_{1}^{n-r}\left(\beta_{1}+(d-c) y\right)^{n-r+1}-\widehat{\gamma}_{1} \beta_{2}^{n-r}\left(\beta_{2}+(d-c) y\right)^{n-r+1}}{\left(a b x^{2}\right)^{n-r}\left(\beta_{1}-\beta_{2}\right)}  \tag{18}\\
B H_{2 n+1}(x, y)=\frac{\widehat{\alpha}_{1} \beta_{1}^{n}\left(\beta_{1}+(d-c) y\right)^{n+1}-\widehat{\gamma}_{1} \beta_{2}^{n}\left(\beta_{2}+(d-c) y\right)^{n+1}}{\left(a b x^{2}\right)^{n}\left(\beta_{1}-\beta_{2}\right)} . \tag{19}
\end{gather*}
$$

By virtue of (17), (18) and (19), we have

$$
\begin{aligned}
& B H_{2(n+r)+1}(x, y) B H_{2(n-r)+1}(x, y)-\left(B H_{2 n+1}(x, y)\right)^{2} \\
& = \\
& \frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\alpha}_{1} \widehat{\gamma}_{1} \beta_{1}^{n} \beta_{2}{ }^{n}\left(\beta_{1}+(d-c) y\right)^{n+1}\left(\beta_{2}+(d-c) y\right)^{n+1}\left[1-\left(\frac{\beta_{1}\left(\beta_{1}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right] \\
& \quad+\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\gamma}_{1} \widehat{\alpha}_{1} \beta_{2}{ }^{n} \beta_{1}^{n}\left(\beta_{2}+(d-c) y\right)^{n+1}\left(\beta_{1}+(d-c) y\right)^{n+1}\left[1-\left(\frac{\beta_{2}\left(\beta_{2}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right]
\end{aligned}
$$

Finally, we get

$$
\begin{aligned}
& B H_{2(n+r)+\xi(i)}(x, y) B H_{2(n-r)+\xi(i)}(x, y)-\left(B H_{2 n+\xi(i)}(x, y)\right)^{2} \\
& =\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\alpha}_{\xi(i)} \widehat{\gamma}_{\xi(i)} \beta_{1}^{n} \beta_{2}^{n}\left(\beta_{1}+(d-c) y\right)^{n+\xi(i)}\left(\beta_{2}+(d-c) y\right)^{n+\xi(i)}\left[1-\left(\frac{\beta_{1}\left(\beta_{1}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right] \\
& \quad+\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\gamma}_{\xi(i)} \widehat{\alpha}_{\xi(i)} \beta_{2}^{n} \beta_{1}^{n}\left(\beta_{2}+(d-c) y\right)^{n+\xi(i)}\left(\beta_{1}+(d-c) y\right)^{n+\xi(i)}\left[1-\left(\frac{\beta_{2}\left(\beta_{2}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)^{r}\right]\right] .
\end{aligned}
$$

Now, we give the Cassini's identity of the bivariate conditional Fibonacci hybrinomial $B H_{n}(x, y)$.
Corollary 1. For $n \geq 0$, we get

$$
\begin{aligned}
& B H_{2(n+1)+\xi(i)}(x, y) B H_{2(n-1)+\xi(i)}(x, y)-\left(B H_{2 n+\xi(i)}(x, y)\right)^{2} \\
& =\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\alpha}_{\xi(i)} \widehat{\gamma}_{\xi(i)} \beta_{1}{ }^{n} \beta_{2}{ }^{n}\left(\beta_{1}+(d-c) y\right)^{n+\xi(i)}\left(\beta_{2}+(d-c) y\right)^{n+\xi(i)}\left[1-\left(\frac{\beta_{1}\left(\beta_{1}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)\right]\right] \\
& \quad+\frac{1}{\left(a b x^{2}\right)^{2 n}\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& \quad \times\left[\widehat{\gamma}_{\xi(i)} \widehat{\alpha}_{\xi(i)} \beta_{2}{ }^{n} \beta_{1}{ }^{n}\left(\beta_{2}+(d-c) y\right)^{n+\xi(i)}\left(\beta_{1}+(d-c) y\right)^{n+\xi(i)}\left[1-\left(\frac{\beta_{2}\left(\beta_{2}+(d-c) y\right)}{\beta_{2}\left(\beta_{2}+(d-c) y\right)}\right)\right]\right] .
\end{aligned}
$$

where $\widehat{\alpha}_{\xi(i)}$ and $\widehat{\gamma}_{\xi(i)}$ are defined in Theorem(2) and $i \in\{0,1\}$.
Proof. Taking $r=1$ in Catalan's identity the proof is completed.

## 3. Generalized Bivariate Conditional Lucas Hybrinomials

In this section we give some identities of the generalized bivariate conditional Lucas hybrinomials. We start with the following definition.
Definition 2. For any four numbers $a, b, c$ and $d$ belonging to $\mathbb{R}-\{0\}$, the generalization of bivariate conditional Fibonacci polynomial is defined as,

$$
L_{n}(x, y)=\left\{\begin{array}{ll}
b x L_{n-1}(x, y)+d y L_{n-2}(x, y), & \text { if } n \text { is even }  \tag{20}\\
a x L_{n-1}(x, y)+c y L_{n-2}(x, y), & \text { if } n \text { is odd }
\end{array} n \geq 2\right.
$$

where $L_{0}(x, y)=2, L_{1}(x, y)=a x$.
Lemma 2. For the generalized bivariate conditional Lucas polynomials $\left\{L_{n}(x, y)\right\}_{n=0}^{\infty}$, we have

$$
\begin{aligned}
& L_{2 n}(x, y)=\left(a b x^{2}+(c+d) y\right) L_{2 n-2}(x, y)-c d y^{2} L_{2 n-4}(x, y) \\
& L_{2 n+1}(x, y)=\left(a b x^{2}+(c+d) y\right) L_{2 n-1}(x, y)-c d y^{2} L_{2 n-3}(x, y)
\end{aligned}
$$

Proof. By using the definition of the generalized bivariate conditional Lucas polynomials, we have

$$
\begin{aligned}
L_{2 n}(x, y) & =\left(b x L_{2 n-1}(x, y)+d y L_{2 n-2}(x, y)\right) \\
& =\left[b x\left(a x L_{2 n-2}(x, y)+c y L_{2 n-3}(x, y)\right)+d y L_{2 n-2}(x, y)\right] \\
& =\left[\left(a b x^{2}+d y\right) L_{2 n-2}(x, y)+c y\left(b x L_{2 n-3}(x, y)\right)\right] \\
& =\left[\left(a b x^{2}+d y\right) L_{2 n-2}(x, y)+c y\left(L_{2 n-2}(x, y)-d y L_{2 n-4}(x, y)\right)\right] \\
& =\left[\left(a b x^{2}+(c+d) y\right) L_{2 n-2}(x, y)-c d y^{2} L_{2 n-4}(x, y)\right] .
\end{aligned}
$$

Similar to above steps, we can obtain

$$
L_{2 n+1}(x, y)=\left(a b x^{2}+(c+d) y\right) L_{2 n-1}(x, y)-c d y^{2} L_{2 n-3}(x, y)
$$

Thus, the proof is completed.

Next we give the generating function for the bivariate conditional Lucas polyno$\operatorname{mial} L_{n}(x, y)$.

Theorem 4. The generating function for the bivariate conditional Lucas polynomial $L_{n}(x, y)$ is

$$
\begin{equation*}
E(t)=\sum_{n=0}^{\infty} L_{n}(x, y) t^{n}=\frac{2+a x t-\left(a b x^{2}+2 c y\right) t^{2}+a d x y t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{21}
\end{equation*}
$$

Proof. We define

$$
\begin{aligned}
& E_{0}(t)=\sum_{n=0}^{\infty} L_{2 n}(x, y) t^{2 n} \\
& E_{1}(t)=\sum_{n=0}^{\infty} L_{2 n+1}(x, y) t^{2 n+1}
\end{aligned}
$$

So that

$$
E(t)=E_{0}(t)+E_{1}(t)
$$

We have

$$
\begin{aligned}
E_{0}(t)= & \sum_{n=0}^{\infty} L_{2 n}(x, y) t^{2 n} \\
= & \sum_{n=0}^{\infty} L_{2 n}(x, y) t^{2 n}=L_{0}(x, y) t^{0}+L_{2}(x, y) t^{2}+\sum_{n=2}^{\infty} L_{2 n}(x, y) t^{2 n} \\
= & L_{0}(x, y)+L_{2}(x, y) t^{2} \\
& +\sum_{n=2}^{\infty}\left[\left(a b x^{2}+(c+d) y\right) L_{2 n-2}(x, y)-c d y^{2} L_{2 n-4}(x, y)\right] t^{2 n} \\
= & L_{0}(x, y)+L_{2}(x, y) t^{2}+\left(a b x^{2}+(c+d) y\right) t^{2} \sum_{n=2}^{\infty} L_{2 n-2}(x, y) t^{2 n-2} \\
& -c d y^{2} t^{4} \sum_{n=2}^{\infty} L_{2 n-4}(x, y) t^{2 n-4} \\
= & 2+\left(a b x^{2}+2 d y\right) t^{2} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2}\left[\sum_{n=2}^{\infty} L_{2 n-2}(x, y) t^{2 n-2}+L_{0}(x, y) t^{0}-L_{0}(x, y) t^{0}\right] \\
& -c d y^{2} t^{4} E_{0}(t)
\end{aligned}
$$

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$$
\begin{aligned}
= & 2+\left(a b x^{2}+2 d y\right) t^{2}+\left(a b x^{2}+(c+d) y\right) t^{2} E_{0}(t) \\
& -2\left(a b x^{2}+(c+d) y\right) t^{2}-c d y^{2} t^{4} E_{0}(t) \\
E_{0}(t)[1- & \left.\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}\right]=2-\left(a b x^{2}+2 c y\right) t^{2}
\end{aligned}
$$

Thus, we get

$$
\begin{equation*}
E_{0}(t)=\frac{2-\left(a b x^{2}+2 c y\right) t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{22}
\end{equation*}
$$

Similarly, we find

$$
\begin{aligned}
E_{1}(t)= & \sum_{n=0}^{\infty} L_{2 n+1}(x, y) t^{2 n+1} \\
= & \sum_{n=0}^{\infty} L_{2 n+1}(x, y) t^{2 n+1}=L_{1}(x, y) t^{1}+L_{3}(x, y) t^{3}+\sum_{n=2}^{\infty} L_{2 n+1}(x, y) t^{2 n+1} \\
= & L_{1}(x, y) t+L_{3}(x, y) t^{3} \\
& +\sum_{n=2}^{\infty}\left[\left(a b x^{2}+(c+d) y\right) L_{2 n-1}(x, y)-c d y^{2} L_{2 n-3}(x, y)\right] t^{2 n+1} \\
= & L_{1}(x, y) t+L_{3}(x, y) t^{3} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2} \sum_{n=2}^{\infty} L_{2 n-1}(x, y) t^{2 n-1} \\
& -c d y^{2} t^{4} \sum_{n=2}^{\infty} L_{2 n-3}(x, y) t^{2 n-3} \\
= & a x t+\left(a^{2} b x^{3}+2 a d x y+a c x y\right) t^{3} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2}\left[\sum_{n=2}^{\infty} L_{2 n-1}(x, y) t^{2 n-1}+L_{1}(x, y) t-L_{1}(x, y) t\right] \\
& -c d y^{2} t^{4} E_{1}(t) \\
= & a x t+\left(a^{2} b x^{3}+2 a d x y+a c x y\right) t^{3}+\left(a b x^{2}+(c+d) y\right) t^{2} E_{1}(t) \\
& -a x\left(a b x^{2}+(c+d) y\right) t^{3}-c d y^{2} t^{4} E_{1}(t) \\
E_{1}(t)[1 & \left.-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}\right]=a x t+a d x y t^{3} .
\end{aligned}
$$

Therefore, we get

$$
\begin{equation*}
E_{1}(t)=\frac{a x t+a d x y t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{23}
\end{equation*}
$$

By virtue of (22) and (23), we can obtain

$$
\begin{aligned}
E(t) & =E_{0}(t)+E_{1}(t) \\
& =\sum_{n=0}^{\infty} L_{n}(x, y) t^{n}=\frac{2+a x t-\left(a b x^{2}+2 c y\right) t^{2}+a d x y t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}}
\end{aligned}
$$

Hence, the proof is completed.
Now we give the Binet formula of the bivariate conditional Lucas polynomial $L_{n}(x, y)$.
Theorem 5. The $n^{\text {th }}$ term of the generalized of bivariate conditional Lucas polynomial $L_{n}(x, y)$ is

$$
\begin{align*}
L_{n}(x, y)= & \frac{(-a x)^{\xi(n)}}{\beta_{1}-\beta_{2}} \\
& \times\left[\left(\xi(n+1) \beta_{1}+(-1)^{\xi(n+1)} \beta_{2}\right)\left(\beta_{2}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}\right. \\
& \quad+\left((-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{2}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}  \tag{24}\\
& \quad-\left(\xi(n+1) \beta_{2}+(-1)^{\xi(n+1)} \beta_{1}\right)\left(\beta_{1}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor} \\
& \left.\quad-\left((-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{1}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}\right]
\end{align*}
$$

where $\beta_{1}$ and $\beta_{2}$ roots of the characteristic equation $\lambda^{2}-\left(a b x^{2}+(c-d) y\right) \lambda-a b d x^{2} y=$ 0 .

Proof. We use the following properties throughout the proof:

- $\beta_{1}+\beta_{2}=a b x^{2}+(c-d) y$
- $\beta_{1} \cdot \beta_{2}=-a b d x^{2} y$
- $\left(\beta_{1}+d y\right)\left(\beta_{2}+d y\right)=c d y^{2}$
- $\left(\beta_{1}+d y\right)\left(a b x^{2}\right)=\beta_{1}\left(\beta_{1}+(d-c) y\right)$
- $\left(\beta_{2}+d y\right)\left(a b x^{2}\right)=\beta_{2}\left(\beta_{2}+(d-c) y\right)$

Note that $\beta_{1}(x, y)=\beta_{1}$ and $\beta_{2}(x, y)=\beta_{2}$. Since $\frac{\beta_{1}+d y}{c d y^{2}}$ and $\frac{\beta_{2}+d y}{c d y^{2}}$ are roots of

$$
1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}=0
$$

If we assume

$$
\begin{aligned}
& E_{0}(t)=\sum_{n=0}^{\infty} L_{2 n}(x, y) t^{2 n} \\
& E_{1}(t)=\sum_{n=0}^{\infty} L_{2 n+1}(x, y) t^{2 n+1}
\end{aligned}
$$

Then,

$$
E(t)=E_{0}(t)+E_{1}(t)
$$

By using Maclaurin's series expansion

$$
\frac{A z-B}{z^{2}-C}=\sum_{n=0}^{\infty} B C^{-n-1} Z^{2 n}-\sum_{n=0}^{\infty} A C^{-n-1} Z^{2 n+1}
$$

and above-mentioned identities, we simplify both $E_{0}(t)$ and $E_{1}(t)$ as follows:

$$
\begin{aligned}
& E_{0}(t)=\frac{1}{c d y^{2}\left(\beta_{1}-\beta_{2}\right)}\left[\frac{2 c d y^{2}-\left(a b x^{2}+2 c y\right) \cdot\left(\beta_{1}+d y\right)}{t^{2}-\left(\frac{\beta_{1}+d y}{c d y^{2}}\right)}\right. \\
& \left.-\frac{2 c d y^{2}-\left(a b x^{2}+2 c y\right)\left(\beta_{2}+y\right)}{t^{2}-\left(\frac{\beta_{2}+d y}{c d y^{2}}\right)}\right] \\
& =\frac{1}{c d y^{2}\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(\left(a b x^{2}+2 c y\right)\left(\beta_{1}+d y\right)-2 c d y^{2}\right)\left(\frac{\beta_{1}+d y}{c d y^{2}}\right)^{-n-1}\right] t^{2 n} \\
& -\frac{1}{c d y^{2}\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(\left(a b x^{2}+2 c y\right)\left(\beta_{2}+d y\right)-2 c d y^{2}\right)\left(\frac{\beta_{2}+d y}{c d y^{2}}\right)^{-n-1}\right] t^{2 n} \\
& =\frac{1}{c d y^{2}\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(\left(a b x^{2}+2 c y\right)\left(\beta_{1}+d y\right)\left(\beta_{2}+d y\right)\right.\right. \\
& \left.\left.-2 c d y^{2}\left(\beta_{2}+d y\right)\right)\left(\beta_{2}+d y\right)^{n}\right] t^{2 n} \\
& -\frac{1}{c d y^{2}\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(\left(a b x^{2}+2 c y\right)\left(\beta_{2}+d y\right)\left(\beta_{1}+d y\right)\right.\right. \\
& \left.\left.-2 c d y^{2}\left(\beta_{1}+d y\right)\right)\left(\beta_{1}+d y\right)^{n}\right] t^{2 n} \\
& =\frac{1}{\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(a b x^{2}-2 \beta_{2}+2 c y-2 d y\right)\left(\beta_{2}+d y\right)^{n}\right. \\
& \left.-\left(a b x^{2}-2 \beta_{1}+2 c y-2 d y\right)\left(\beta_{1}+d y\right)^{n}\right] t^{2 n} \\
& =\frac{1}{\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(\beta_{1}-\beta_{2}-(d-c) y\right)\left(\beta_{2}+d y\right)^{n}\right. \\
& \left.-\left(\beta_{2}-\beta_{1}-(d-c) y\right)\left(\beta_{1}+d y\right)^{n}\right] t^{2 n} .
\end{aligned}
$$

We solve $E_{1}(t)$ with the same approach used in $E_{0}(t)$ and we get the value of

$$
E_{1}(t)=\frac{-a x}{\left(\beta_{1}-\beta_{2}\right)} \sum_{n=0}^{\infty}\left[\left(\beta_{2}+2 d y\right)\left(\beta_{2}+d y\right)^{n}-\left(\beta_{1}+2 d y\right)\left(\beta_{1}+d y\right)^{n}\right] t^{2 n+1}
$$

We know that $E(t)=E_{0}(t)+E_{1}(t)$. So we find

$$
\begin{aligned}
& E(t)=\sum_{n=0}^{\infty} \frac{(-a x)^{\xi(n)}[ }{\beta_{1}-\beta_{2}}\left[\begin{array}{l}
( \\
\xi(n
\end{array}+1\right) \beta_{1}+(-1)^{\xi(n+1)} \beta_{2} \\
&\left.\quad+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{2}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor} \\
& \quad-\left(\xi(n+1) \beta_{2}+(-1)^{\xi(n+1)} \beta_{1}\right. \\
&\left.\left.\quad+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{1}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}\right] .
\end{aligned}
$$

Thus,

$$
\begin{aligned}
L_{n}(x, y)=\frac{(-a x)^{\xi(n)}}{\beta_{1}-\beta_{2}}[(\xi(n & +1) \beta_{1}+(-1)^{\xi(n+1)} \beta_{2} \\
& \left.+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{2}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor} \\
& -\left(\xi(n+1) \beta_{2}+(-1)^{\xi(n+1)} \beta_{1}\right. \\
& \left.\left.+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{1}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}\right]
\end{aligned}
$$

In the following definition, we give bivariate conditional Lucas Hybrinomials.
Definition 3. For any variable $x, y$ and nonzero real numbers $a, b, c$ and $d$, we have

$$
\begin{equation*}
L H_{n}(x, y)=L_{n}(x, y)+\boldsymbol{i} L_{n+1}(x, y)+\varepsilon L_{n+2}(x, y)+\boldsymbol{h} L_{n+3}(x, y) \tag{25}
\end{equation*}
$$

where $L_{n}(x, y)$ was given in and the initial conditions are with $L H_{0}(x, y)=$ $2+\boldsymbol{i} a x+\varepsilon\left(a b x^{2}+2 d y\right)+\boldsymbol{h}\left(a^{2} b x^{3}+2 a d x y+a c x y\right)$ and $L H_{1}(x, y)=a x+\boldsymbol{i}\left(a b x^{2}+\right.$ $2 d y)++\varepsilon\left(a^{2} b x^{3}+2 a d x y+a c x y\right)+\boldsymbol{h}\left(a^{2} b^{2} x^{4}+2 b c d x^{2} y+a b c x^{2} y+a b d x^{2} y+2 d^{2} y^{2}\right)$.

We can see from the following table that the generalized bivariate conditional Fibonacci hybrinomials are the generalization of many works for different values of $a, b, c$ and $d$.

Table 3. The generalized bivariate conditional Lucas hybrinomials

| $\boldsymbol{a}$ | $\boldsymbol{b}$ | $\boldsymbol{c}$ | $\boldsymbol{d}$ | Generalized Bivariate Conditional Lucas Hybrinomials |
| :---: | :---: | :---: | :---: | :---: |
| 1 | 1 | 1 | 1 | Bivariate Lucas Hybrinomials |
| $a$ | $b$ | 1 | 1 | Bivariate Conditional Lucas Hybrinomials |
| 2 | 2 | 1 | 1 | Bivariate Pell Lucas Hybrinomials |
| 1 | 1 | 2 | 2 | Bivariate Jacobsthal Lucas Hybrinomials |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |

Lemma 3. For the generalized bivariate conditional Lucas hybrinomials $\left\{L H_{n}(x, y)\right\}_{n=0}^{\infty}$, we have

$$
\begin{aligned}
& L H_{2 n}(x, y)=\left(a b x^{2}+(c+d) y\right) L H_{2 n-2}(x, y)-c d y^{2} L H_{2 n-4}(x, y) \\
& L H_{2 n+1}(x, y)=\left(a b x^{2}+(c+d) y\right) L H_{2 n-1}(x, y)-c d y^{2} L H_{2 n-3}(x, y) .
\end{aligned}
$$

Proof. By using the definition of the generalized bivariate conditional Lucas hybrinomials, we obtain

$$
\begin{aligned}
L H_{2 n}(x, y)= & L_{2 n}(x, y)+\mathbf{i} L_{2 n+1}(x, y)+\varepsilon L_{2 n+2}(x, y)+\mathbf{h} L_{2 n+3}(x, y) \\
= & \left(b x L_{2 n-1}(x, y)+d y L_{2 n-2}(x, y)\right)+\mathbf{i}\left(a x L_{2 n}(x, y)+c y L_{2 n-1}(x, y)\right) \\
& +\varepsilon\left(b x L_{2 n+1}(x, y)+d y L_{2 n}(x, y)\right)+\mathbf{h}\left(a x L_{2 n+2}(x, y)+c y L_{2 n+1}(x, y)\right) \\
= & {\left[b x\left(a x L_{2 n-2}(x, y)+c y L_{2 n-3}(x, y)\right)+d y L_{2 n-2}(x, y)\right] } \\
& +\mathbf{i}\left[a x\left(b x L_{2 n-1}(x, y)+d y L_{2 n-2}(x, y)\right)+c y L_{2 n-1}(x, y)\right] \\
& \left.+\varepsilon\left[b x\left(a x L_{2 n}(x, y)+c y L_{2 n-1}(x, y)\right)+d y L_{2 n}(x, y)\right)\right] \\
& +\mathbf{h}\left[a x\left(b x L_{2 n+1}(x, y)+d y L_{2 n}(x, y)\right)+c y L_{2 n+1}(x, y)\right] \\
= & {\left[\left(a b x^{2}+d y\right) L_{2 n-2}(x, y)+c y\left(b x L_{2 n-3}(x, y)\right)\right] } \\
& +\mathbf{i}\left[\left(a b x^{2}+c y\right) L_{2 n-1}(x, y)+d y\left(a x L_{2 n-2}(x, y)\right)\right] \\
& +\varepsilon\left[\left(a b x^{2}+d y\right) L_{2 n}(x, y)+c y\left(b x L_{2 n-1}(x, y)\right)\right] \\
& +\mathbf{h}\left[\left(a b x^{2}+c y\right) L_{2 n+1}(x, y)+d y\left(a x L_{2 n}(x, y)\right)\right] \\
= & {\left[\left(a b x^{2}+d y\right) L_{2 n-2}(x, y)+c y\left(L_{2 n-2}(x, y)-d y L_{2 n-4}(x, y)\right)\right] } \\
& +\mathbf{i}\left[\left(a b x^{2}+c y\right) L_{2 n-1}(x, y)+d y\left(L_{2 n-1}(x, y)-c y L_{2 n-3}(x, y)\right)\right] \\
& +\varepsilon\left[\left(a b x^{2}+d y\right) L_{2 n}(x, y)+c y\left(L_{2 n}(x, y)-d y L_{2 n-2}(x, y)\right)\right] \\
& +\mathbf{h}\left[\left(a b x^{2}+c y\right) L_{2 n+1}(x, y)+d y\left(L_{2 n+1}(x, y)-c y L_{2 n-1}(x, y)\right)\right] \\
= & {\left[\left(a b x^{2}+(c+d) y\right) L_{2 n-2}(x, y)-c d y^{2} L_{2 n-4}(x, y)\right] } \\
& +\mathbf{i}\left[\left(a b x^{2}+(c+d) y\right) L_{2 n-1}(x, y)-c d y^{2} L_{2 n-3}(x, y)\right] \\
& +\varepsilon\left[\left(a b x^{2}+(c+d) y\right) L_{2 n}(x, y)-c d y^{2} L_{2 n-2}(x, y)\right] \\
& +\mathbf{h}\left[\left(a b x^{2}+(c+d) y\right) L_{2 n+1}(x, y)-c d y L_{2 n-1}(x, y)\right]
\end{aligned}
$$

$$
\begin{aligned}
= & \left(a b x^{2}+(c+d) y\right)\left[L_{2 n-2}(x, y)+\mathbf{i} L_{2 n-1}(x, y)+\varepsilon L_{2 n}(x, y)+\mathbf{h} L_{2 n+1}(x, y)\right] \\
& -c d y^{2}\left[L_{2 n-4}(x, y)+\mathbf{i} L_{2 n-3}(x, y)+\varepsilon L_{2 n-2}(x, y)+\mathbf{h} L_{2 n-1}(x, y)\right] \\
= & \left(a b x^{2}+(c+d) y\right) L H_{2 n-2}(x, y)-c d y^{2} L H_{2 n-4}(x, y)
\end{aligned}
$$

Similar to above, we can obtain

$$
L H_{2 n+1}(x, y)=\left(a b x^{2}+(c+d) y\right) L H_{2 n-1}(x, y)-c d y^{2} L H_{2 n-3}(x, y)
$$

Thus, the proof is completed.

Next we give the generating function of the bivariate conditional Lucas hybrinomial $L H_{n}(x, y)$.

Theorem 6. The generating function for the bivariate conditional Lucas hybrinomial $L H_{n}(x, y)$ is

$$
\begin{aligned}
\Omega(t)= & \sum_{n=0}^{\infty} L H_{n}(x, y) t^{n} \\
= & \frac{L H_{0}(x, y)+L H_{1}(x, y) t+\left[L H_{2}(x, y)-\left(a b x^{2}+(c+d) y\right) L H_{0}(x, y)\right] t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \\
& +\frac{\left[L H_{3}(x, y)-\left(a b x^{2}+(c+d) y\right) L H_{1}(x, y)\right] t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}}
\end{aligned}
$$

Proof. We define

$$
\begin{aligned}
& \Omega_{0}(t)=\sum_{n=0}^{\infty} L H_{2 n}(x, y) t^{2 n} \\
& \Omega_{1}(t)=\sum_{n=0}^{\infty} L H_{2 n+1}(x, y) t^{2 n+1}
\end{aligned}
$$

So that

$$
\Omega(t)=\Omega_{0}(t)+\Omega_{1}(t)
$$

We have

$$
\begin{aligned}
\Omega_{0}(t)= & \sum_{n=0}^{\infty} L H_{2 n}(x, y) t^{2 n} \\
= & \sum_{n=0}^{\infty} L H_{2 n}(x, y) t^{2 n}=L H_{0}(x, y) t^{0}+L H_{2}(x, y) t^{2}+\sum_{n=2}^{\infty} L H_{2 n}(x, y) t^{2 n} \\
= & L H_{0}(x, y)+L H_{2}(x, y) t^{2} \\
& +\sum_{n=2}^{\infty}\left[\left(a b x^{2}+(c+d) y\right) L H_{2 n-2}(x, y)-c d y^{2} L H_{2 n-4}(x, y)\right] t^{2 n} \\
= & L H_{0}(x, y)+L H_{2}(x, y) t^{2}+\left(a b x^{2}+(c+d) y\right) t^{2} \sum_{n=2}^{\infty} L H_{2 n-2}(x, y) t^{2 n-2} \\
& -c d y^{2} t^{4} \sum_{n=2}^{\infty} L H_{2 n-4}(x, y) t^{2 n-4} \\
= & L H_{0}(x, y)+L H_{2}(x, y) t^{2} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2}\left[\sum_{n=2}^{\infty} L H_{2 n-2}(x, y) t^{2 n-2}+L H_{0}(x, y) t^{0}-L H_{0}(x, y) t^{0}\right] \\
& -c d y^{2} t^{4} \Omega_{0}(t) \\
= & L H_{0}(x, y)+L H_{2}(x, y) t^{2}+\left(a b x^{2}+(c+d) y\right) t^{2} \Omega_{0}(t) \\
& -\left(a b x^{2}+(c+d) y\right) t^{2} L H_{0}(x, y)-c d y^{2} t^{4} \Omega_{0}(t) .
\end{aligned}
$$

Thus, we get

$$
\begin{equation*}
\Omega_{0}(t)=\frac{L H_{0}(x, y)+\left(L H_{2}(x, y)-\left(a b x^{2}+(c+d) y\right) L H_{0}(x, y)\right) t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{26}
\end{equation*}
$$

Similarly, we find

$$
\begin{aligned}
\Omega_{1}(t)= & \sum_{n=0}^{\infty} L H_{2 n+1}(x, y) t^{2 n+1} \\
= & \sum_{n=0}^{\infty} L H_{2 n+1}(x, y) t^{2 n+1}=L H_{1}(x, y) t+L H_{3}(x, y) t^{3} \\
& +\sum_{n=2}^{\infty} L H_{2 n+1}(x, y) t^{2 n+1} \\
= & L H_{1}(x, y) t+L H_{3}(x, y) t^{3} \\
& +\sum_{n=2}^{\infty}\left[\left(a b x^{2}+(c+d) y\right) L H_{2 n-1}(x, y)-c d y^{2} L H_{2 n-3}(x, y)\right] t^{2 n+1}
\end{aligned}
$$

$$
\begin{aligned}
= & L H_{1}(x, y) t+L H_{3}(x, y) t^{3} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2} \sum_{n=2}^{\infty} L H_{2 n-1}(x, y) t^{2 n-1}-c d y^{2} t^{4} \sum_{n=2}^{\infty} L H_{2 n-3}(x, y) t^{2 n-3} \\
= & L H_{1}(x, y) t+L H_{3}(x, y) t^{3} \\
& +\left(a b x^{2}+(c+d) y\right) t^{2}\left[\sum_{n=2}^{\infty} L H_{2 n-1}(x, y) t^{2 n-1}+L H_{1}(x, y) t-L H_{1}(x, y) t\right] \\
& -c d y^{2} t^{4} \Omega_{1}(t) \\
= & L H_{1}(x, y) t+L H_{3}(x, y) t^{3}+\left(a b x^{2}+(c+d) y\right) t^{2} \Omega_{1}(t) \\
& -\left(a b x^{2}+(c+d) y\right) t^{3} L H_{1}(x, y)-c d y^{2} t^{4} \Omega_{1}(t)
\end{aligned}
$$

Therefore, we get

$$
\begin{equation*}
\Omega_{1}(t)=\frac{L H_{1}(x, y) t+\left(L H_{3}(x, y)-\left(a b x^{2}+(c+d) y\right) L H_{1}(x, y)\right) t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \tag{27}
\end{equation*}
$$

By virtue of (26) and (27), we can obtain

$$
\begin{aligned}
\Omega(t)= & \Omega_{0}(t)+\Omega_{1}(t) \\
= & \sum_{n=0}^{\infty} L H_{n}(x, y) t^{n} \\
= & \frac{L H_{0}(x, y)+L H_{1}(x, y) t+\left[L H_{2}(x, y)-\left(a b x^{2}+(c+d) y\right) L H_{0}(x, y)\right] t^{2}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}} \\
& +\frac{\left[L H_{3}(x, y)-\left(a b x^{2}+(c+d) y\right) L H_{1}(x, y)\right] t^{3}}{1-\left(a b x^{2}+(c+d) y\right) t^{2}+c d y^{2} t^{4}}
\end{aligned}
$$

Hence, the proof is completed.

Now we give the Binet formula of the bivariate conditional Lucas hybrinomial $L H_{n}(x, y)$.

Theorem 7. The $n^{\text {th }}$ term of the generalized of bivariate conditional Lucas hybrinomial $L H_{n}(x, y)$ is

$$
\begin{equation*}
L H_{n}(x, y)=\frac{\widehat{\omega}_{\xi(n)}\left(\beta_{2}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}-\widehat{\sigma}_{\xi(n)}\left(\beta_{1}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}}{\beta_{1}-\beta_{2}} \tag{28}
\end{equation*}
$$

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where $\beta_{1}$ and $\beta_{2}$ roots of the characteristic equation $\lambda^{2}-\left(a b x^{2}+(c-d) y\right) \lambda-a b d x^{2} y=$ 0 . Also,

$$
\begin{aligned}
\widehat{\omega}_{\xi(n)}= & (-a x)^{\xi(n)}\left(\xi(n+1) \beta_{1}+(-1)^{\xi(n+1)} \beta_{2}+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right) \\
& +i(-a x)^{\xi(n+1)}\left(\xi(n) \beta_{1}+(-1)^{\xi(n)} \beta_{2}+(-1)^{\xi(n)}(2)^{\xi(n+1)} d y+\xi(n) c y\right)\left(\beta_{2}+d y\right)^{\xi(n)} \\
& +\varepsilon(-a x)^{\xi(n)}\left(\xi(n+1) \beta_{1}+(-1)^{\xi(n+1)} \beta_{2}+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{2}+d y\right) \\
& +h(-a x)^{\xi(n+1)}\left(\xi(n) \beta_{1}+(-1)^{\xi(n)} \beta_{2}+(-1)^{\xi(n)}(2)^{\xi(n+1)} d y+\xi(n) c y\right)\left(\beta_{2}+d y\right)^{\xi(n)+1}
\end{aligned}
$$

and

$$
\begin{aligned}
\widehat{\sigma}_{\xi(n)}= & (-a x)^{\xi(n)}\left(\xi(n+1) \beta_{2}+(-1)^{\xi(n+1)} \beta_{1}+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right) \\
& +i(-a x)^{\xi(n+1)}\left(\xi(n) \beta_{2}+(-1)^{\xi(n)} \beta_{1}+(-1)^{\xi(n)}(2)^{\xi(n+1)} d y+\xi(n) c y\right)\left(\beta_{1}+d y\right)^{\xi(n)} \\
& +\varepsilon(-a x)^{\xi(n)}\left(\xi(n+1) \beta_{2}+(-1)^{\xi(n+1)} \beta_{1}+(-1)^{\xi(n+1)}(2)^{\xi(n)} d y+\xi(n+1) c y\right)\left(\beta_{1}+d y\right) \\
& +h(-a x)^{\xi(n+1)}\left(\xi(n) \beta_{2}+(-1)^{\xi(n)} \beta_{1}+(-1)^{\xi(n)}(2)^{\xi(n+1)} d y+\xi(n) c y\right)\left(\beta_{1}+d y\right)^{\xi(n)+1} .
\end{aligned}
$$

Proof. Firstly, by using (24), we have

$$
\begin{aligned}
L H_{2 n}(x, y)= & L_{2 n}(x, y)+\mathbf{i} L_{2 n+1}(x, y)+\varepsilon L_{2 n+2}(x, y)+\mathbf{h} L_{2 n+3}(x, y) \\
=\frac{\left(\beta_{2}+d y\right)^{n}}{\beta_{1}-\beta_{2}}[ & \left(\beta_{1}-\beta_{2}-d y+c y\right)+\mathbf{i}(-a x)\left(\beta_{2}+2 d y\right) \\
& +\varepsilon\left(\beta_{1}-\beta_{2}-d y+c y\right)\left(\beta_{2}+d y\right) \\
& \left.+\mathbf{h}(-a x)\left(\beta_{2}+2 d y\right)\left(\beta_{2}+d y\right)\right] \\
-\frac{\left(\beta_{1}+d y\right)^{n}}{\beta_{1}-\beta_{2}} & {\left[\left(\beta_{2}-\beta_{1}-d y+c y\right)+\mathbf{i}(-a x)\left(\beta_{1}+2 d y\right)\right.} \\
& +\varepsilon\left(\beta_{2}-\beta_{1}-d y+c y\right)\left(\beta_{1}+d y\right) \\
& \left.+\mathbf{h}(-a x)\left(\beta_{1}+2 d y\right)\left(\beta_{1}+d y\right)\right] .
\end{aligned}
$$

Here, we choose the $\widehat{\omega}_{0}$ and $\widehat{\sigma}_{0}$ as follows;

$$
\begin{aligned}
\widehat{\omega}_{0}= & {\left[\left(\beta_{1}-\beta_{2}-d y+c y\right)+\mathbf{i}(-a x)\left(\beta_{2}+2 d y\right)+\varepsilon\left(\beta_{1}-\beta_{2}-d y+c y\right)\left(\beta_{2}+d y\right)\right.} \\
& \left.+\mathbf{h}(-a x)\left(\beta_{2}+2 d y\right)\left(\beta_{2}+d y\right)\right] \\
\widehat{\sigma}_{0}= & {\left[\left(\beta_{2}-\beta_{1}-d y+c y\right)+\mathbf{i}(-a x)\left(\beta_{1}+2 d y\right)+\varepsilon\left(\beta_{2}-\beta_{1}-d y+c y\right)\left(\beta_{1}+d y\right)\right.} \\
& \left.+\mathbf{h}(-a x)\left(\beta_{1}+2 d y\right)\left(\beta_{1}+d y\right)\right] .
\end{aligned}
$$

Finally, the following equation is obtained:

$$
\begin{equation*}
L H_{2 n}(x, y)=\frac{\widehat{\omega}_{0}\left(\beta_{2}+d y\right)^{n}-\widehat{\sigma}_{0}\left(\beta_{1}+d y\right)^{n}}{\beta_{1}-\beta_{2}} . \tag{29}
\end{equation*}
$$

In similar way, by using (24), we have

$$
\begin{aligned}
L H_{2 n+1}(x, y)=L_{2 n+1}(x, y) & +\mathbf{i} L_{2 n+2}(x, y)+\varepsilon L_{2 n+3}(x, y)+\mathbf{h} L_{2 n+4}(x, y) \\
=\frac{\left(\beta_{2}+d y\right)^{n}}{\beta_{1}-\beta_{2}}[ & (-a x)\left(\beta_{2}+2 d y\right)+\mathbf{i}\left(\beta_{1}-\beta_{2}-d y+c y\right)\left(\beta_{2}+d y\right) \\
& +\varepsilon(-a x)\left(\beta_{2}+2 d y\right)\left(\beta_{2}+d y\right) \\
& \left.+\mathbf{h}\left(\beta_{1}-\beta_{2}-d y+c y\right)\left(\beta_{2}+d y\right)^{2}\right] \\
-\frac{\left(\beta_{1}+d y\right)^{n}}{\beta_{1}-\beta_{2}} & {\left[(-a x)\left(\beta_{1}+2 d y\right)+\mathbf{i}\left(\beta_{2}-\beta_{1}-d y+c y\right)\left(\beta_{1}+d y\right)\right.} \\
& +\varepsilon(-a x)\left(\beta_{1}+2 d y\right)\left(\beta_{1}+d y\right) \\
& \left.+\mathbf{h}\left(\beta_{2}-\beta_{1}-d y+c y\right)\left(\beta_{1}+d y\right)^{2}\right] .
\end{aligned}
$$

Here, we choose the $\widehat{\omega}_{1}$ and $\widehat{\sigma}_{1}$ as follows;

$$
\begin{aligned}
\widehat{\omega}_{1}=[ & (-a x)\left(\beta_{2}+2 d y\right)+\mathbf{i}\left(\beta_{1}-\beta_{2}-d y+c y\right)\left(\beta_{2}+d y\right) \\
& \left.\quad+\varepsilon(-a x)\left(\beta_{2}+2 d y\right)\left(\beta_{2}+d y\right)+\mathbf{h}\left(\beta_{1}-\beta_{2}-d y+c y\right)\left(\beta_{2}+d y\right)^{2}\right] \\
\widehat{\sigma}_{1}=[ & (-a x)\left(\beta_{1}+2 d y\right)+\mathbf{i}\left(\beta_{2}-\beta_{1}-d y+c y\right)\left(\beta_{1}+d y\right) \\
& \left.+\varepsilon(-a x)\left(\beta_{1}+2 d y\right)\left(\beta_{1}+d y\right)+\mathbf{h}\left(\beta_{2}-\beta_{1}-d y+c y\right)\left(\beta_{1}+d y\right)^{2}\right] .
\end{aligned}
$$

Finally, the following equation is obtained.

$$
\begin{equation*}
L H_{2 n+1}(x, y)=\frac{\widehat{\omega}_{1}\left(\beta_{2}+d y\right)^{n}-\widehat{\sigma}_{1}\left(\beta_{1}+d y\right)^{n}}{\beta_{1}-\beta_{2}} \tag{30}
\end{equation*}
$$

By virtue of (29) and (30), we can obtain the following equation

$$
L H_{n}(x, y)=\frac{\widehat{\omega}_{\xi(n)}\left(\beta_{2}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}-\widehat{\sigma}_{\xi(n)}\left(\beta_{1}+d y\right)^{\left\lfloor\frac{n}{2}\right\rfloor}}{\beta_{1}-\beta_{2}}
$$

where $\beta_{1}$ and $\beta_{2}$ roots of the characteristic equation $\lambda^{2}-\left(a b x^{2}+(c-d) y\right) \lambda-$ $a b d x^{2} y=0$.

Now, we give the Catalan's identity of the bivariate conditional Lucas hybrinomial $L H_{n}(x, y)$.
Theorem 8. For any integers $n$ and $r$ and $n \geqslant r \geqslant 0, r \geq 0$, we have

$$
\begin{aligned}
L H_{2(n+r)+\xi(i)}(x, y) L & H_{2(n-r)+\xi(i)}(x, y)-\left(L H_{2 n+\xi(i)}(x, y)\right)^{2} \\
= & \frac{\widehat{\omega}_{\xi(i)} \widehat{\sigma}_{\xi(i)}\left(\beta_{2}+d y\right)^{n}\left(\beta_{1}+d y\right)^{n}\left[1-\left(\frac{\beta_{2}+d y}{\beta_{1}+d y}\right)^{r}\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& +\frac{\widehat{\sigma}_{\xi(i)} \widehat{\omega}_{\xi(i)}\left(\beta_{1}+d y\right)^{n}\left(\beta_{2}+d y\right)^{n}\left[1-\left(\frac{\beta_{1}+d y}{\beta_{2}+d y}\right)^{r}\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}} .
\end{aligned}
$$

where $\widehat{\omega}_{\xi(i)}$ and $\widehat{\sigma}_{\xi(i)}$ are defined in Theorem (7) and $i \in\{0,1\}$.

Proof. In order to prove Catalan's identity, we will examine two different cases.
Case $i=0$ :

$$
\begin{align*}
& L H_{2(n+r)}(x, y)=\frac{\widehat{\omega}_{\xi(2 n+2 r)}\left(\beta_{2}+d y\right)^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor}-\widehat{\sigma}_{\xi(2 n+2 r)}\left(\beta_{1}+d y\right)^{\left\lfloor\frac{2 n+2 r}{2}\right\rfloor}}{\beta_{1}-\beta_{2}} \\
&=\frac{\widehat{\omega}_{0}\left(\beta_{2}+d y\right)^{n+r}-\widehat{\sigma}_{0}\left(\beta_{1}+d y\right)^{n+r}}{\beta_{1}-\beta_{2}}  \tag{31}\\
& L H_{2(n-r)}(x, y)=\frac{\widehat{\omega}_{\xi(2 n-2 r)}\left(\beta_{2}+d y\right)^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor}-\widehat{\sigma}_{\xi(2 n-2 r)}\left(\beta_{1}+d y\right)^{\left\lfloor\frac{2 n-2 r}{2}\right\rfloor}}{\beta_{1}-\beta_{2}}  \tag{32}\\
&=\frac{\widehat{\omega}_{0}\left(\beta_{2}+d y\right)^{n-r}-\widehat{\sigma}_{0}\left(\beta_{1}+d y\right)^{n-r}}{\beta_{1}-\beta_{2}} \\
& L H_{2 n}(x, y)=\frac{\widehat{\omega}_{\xi(2 n)}\left(\beta_{2}+d y\right)^{\left\lfloor\frac{2 n}{2}\right\rfloor}-\widehat{\sigma}_{\xi(2 n)}\left(\beta_{1}+d y\right)^{\left\lfloor\frac{2 n}{2}\right\rfloor}}{\beta_{1}-\beta_{2}}  \tag{33}\\
&=\frac{\widehat{\omega}_{0}\left(\beta_{2}+d y\right)^{n}-\widehat{\sigma}_{0}\left(\beta_{1}+d y\right)^{n}}{\beta_{1}-\beta_{2}}
\end{align*}
$$

By virtue of (31), (32) and (33), we have

$$
\begin{aligned}
L H_{2(n+r)}(x, y) L & H_{2(n-r)}(x, y)-\left(L H_{2 n}(x, y)\right)^{2} \\
= & \frac{\widehat{\omega}_{0} \widehat{\sigma}_{0}\left(\beta_{2}+d y\right)^{n}\left(\beta_{1}+d y\right)^{n}\left[1-\left(\frac{\beta_{2}+d y}{\beta_{1}+d y}\right)^{r}\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& +\frac{\widehat{\sigma}_{0} \widehat{\omega}_{0}\left(\beta_{1}+d y\right)^{n}\left(\beta_{2}+d y\right)^{n}\left[1-\left(\frac{\beta_{1}+d y}{\beta_{2}+d y}\right)^{r}\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}} .
\end{aligned}
$$

Case $i=1$ :

$$
\begin{align*}
L H_{2(n+r)+1}(x, y) & =\frac{\widehat{\omega}_{1}\left(\beta_{2}+d y\right)^{n+r}-\widehat{\sigma}_{1}\left(\beta_{1}+d y\right)^{n+r}}{\beta_{1}-\beta_{2}}  \tag{34}\\
L H_{2(n-r)+1}(x, y) & =\frac{\widehat{\omega}_{1}\left(\beta_{2}+d y\right)^{n-r}-\widehat{\sigma}_{1}\left(\beta_{1}+d y\right)^{n-r}}{\beta_{1}-\beta_{2}}  \tag{35}\\
L H_{2 n+1}(x, y) & =\frac{\widehat{\omega}_{1}\left(\beta_{2}+d y\right)^{n}-\widehat{\sigma}_{1}\left(\beta_{1}+d y\right)^{n}}{\beta_{1}-\beta_{2}} \tag{36}
\end{align*}
$$

By virtue of (34), (35) and (36), we have

$$
\begin{aligned}
L H_{2(n+r)+1}(x, y) L & H_{2(n-r)+1}(x, y)-\left(L H_{2 n+1}(x, y)\right)^{2} \\
= & \frac{\widehat{\omega}_{1} \widehat{\sigma}_{1}\left(\beta_{2}+d y\right)^{n}\left(\beta_{1}+d y\right)^{n}\left[1-\left(\frac{\beta_{2}+d y}{\beta_{1}+d y}\right)^{r}\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& +\frac{\widehat{\sigma}_{1} \widehat{\omega}_{1}\left(\beta_{1}+d y\right)^{n}\left(\beta_{2}+d y\right)^{n}\left[1-\left(\frac{\beta_{1}+d y}{\beta_{2}+d y}\right)^{r}\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}}
\end{aligned}
$$

Thus, the proof is completed.
Now, we give the Cassini's identity of the bivariate conditional Lucas hybrinomial $L H_{n}(x, y)$.

Corollary 2. For $n \geq 0$, we get

$$
\begin{aligned}
L H_{2(n+1)+\xi(i)}(x, y) L & H_{2(n-1)+\xi(i)}(x, y)-\left(L H_{2 n+\xi(i)}(x, y)\right)^{2} \\
= & \frac{\widehat{\sigma}_{\xi(i)} \widehat{\omega}_{\xi(i)}\left(\beta_{1}+d y\right)^{n}\left(\beta_{2}+d y\right)^{n}\left[1-\left(\frac{\beta_{1}+d y}{\beta_{2}+d y}\right)\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}} \\
& +\frac{\widehat{\sigma}_{\xi(i)} \widehat{\omega}_{\xi(i)}\left(\beta_{1}+d y\right)^{n}\left(\beta_{2}+d y\right)^{n}\left[1-\left(\frac{\beta_{1}+d y}{\beta_{2}+d y}\right)\right]}{\left(\beta_{1}-\beta_{2}\right)^{2}}
\end{aligned}
$$

where $\widehat{\omega}_{\xi(i)}$ and $\widehat{\sigma}_{\xi(i)}$ are defined in Theorem(7) and $i \in\{0,1\}$.
Proof. Taking $r=1$ in Catalan's identity the proof is completed.

## 4. Conclusion

The Fibonacci and Lucas numbers are well-known numbers, which have been studied by many researchers for years. These numbers arise in the applications of mathematics, computer science, physics, biology and statistics 9 . In this paper, by combining the Fibonacci and Lucas numbers with hybrid numbers, we present the generalized bivariate conditional Fibonacci and Lucas hybrinomials which are generalization of many works in the literature. Moreover, we derive many properties of generalized bivariate conditional Fibonacci and Lucas hybrinomials such as Binet's formulas, Catalan's identity, Cassini's identity of the hybrinomials.

Author Contribution Statements All authors contributed equally to this manuscript and they read and approved the final manuscript.

Declaration of Competing Interests The authors declare that they have no competing interest.

Acknowledgment This study is a part of the second author's Master Thesis.

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Volume 73, Number 1, Pages 64-75 (2024)
DOI:10.31801/cfsuasmas. 1170867
ISSN 1303-5991 E-ISSN 2618-6470
Research Article; Received: September 4, 2022; Accepted: October 23, 2023

# TRAJECTORY CURVES AND SURFACES: A NEW PERSPECTIVE VIA PROJECTIVE GEOMETRIC ALGEBRA 

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#### Abstract

The aim of this work is to define quaternion curves and surfaces and their conjugates via operators in Euclidean projective geometric algebra (EPGA). In this space, quaternions were obtained by the geometric product of vector fields. New vector fields, which we call trajectory curves and surfaces, were obtained by using this new quaternion operator. Moreover, dual quaternion curves are determined by a similar method and then their generated motion is studied. Illustrative examples are given.


## 1. Introduction

Understanding what complex numbers indicate geometrically has always been a matter of curiosity. Since the problem of finding the roots of a quadratic equation, we use a combination of a complex unit and real numbers, or their ordered binary representation, to show complex numbers. So what does this imaginary part show geometrically? Common usage is an axis orthogonal to the real axis. Thus, it shows the 2-dimensional real space in binary terms. However, complex numbers seem to contain more than that.

While working on the algebra of complex numbers of the form $a+b i$, Argand noticed that when the complex number is multiplied by the imaginary unit $i$, i.e. $i(a+b i)=-b+a i$ represents the rotation of this point, a geometric indicator of the complex number, about the origin in the plane by $90^{\circ} \sqrt{1,2}$. Hamilton thought that this rotational property of complex numbers might also have a counterpart in 3 -dimensional space. So he predicted that an ordered 3 with two complex units could show the rotation in 3-dimensional space. However, it was not that easy to establish the algebra. An undesirable complex expression was coming from the product of two triplets. He used a combination of three imaginary numbers and

[^4]a real number to overcome this problem, and thus multiplication must be closed. Therefore, he invented the quaternions 3:

A (real) quaternion is represented by $q=a+b \mathbf{i}+c \mathbf{j}+d \mathbf{k}$ (general assumption) where $a, b, c, d \in \mathbb{R}$ and the conditions for units

$$
\mathbf{i}^{2}=\mathbf{j}^{2}=\mathbf{k}^{2}=\mathbf{i} \mathbf{j} \mathbf{k}=-1
$$

held.
Thanks to quaternion algebra, one can obtain information about the rotation of any rigid body in space. Especially, in computer graphics, there are two options to rotate an object: use a rotation matrix (3x3 matrix) or a quaternion. Besides, there are some advantages to using quaternion like memory space, speed, and not showing gimbal locks as matrices do.

One can assume coefficients of the quaternion as dual numbers and then get the dual quaternion, i.e. $A, B, C, D \in \mathbb{D} 4 \cdot 6$.
Dual quaternions store both rotation and translation information about a rigid body in space. A similar operation can be done with $4 \times 4$ matrices. It is used quite often in the theory of mechanisms. This type of quaternion pair (called a dual quaternion) stores the same type of information but in two different quaternions. While the dual part of the dual quaternion represents translation information, the real one represents information about rotation. When mentioned together with these features, it is understood that it is a tool for software in robotics. Rather than its usual general definitions, it will be given here in accordance with projective geometric algebra.

Geometric algebra has very useful content to determine objects, especially complex ones geometrically, and transform one to another in space. Generally, geometric algebra over $n$-dimensional vector space is represented by the set $\mathcal{G}(n, 0)$ where $n$ is the dimension of vector space and the second part is the grade of the inner product. Since the vector space studied here is real, the grade of the inner product is 0. Basically, geometric algebra allows us to symbolize scalars, vectors, areas, and volumes using a simple and consistent notation. Such items are closed under the algebraic operation. It is not difficult to see that many variables of this type are closed under addition. However, the product is somewhat unusual. So these expressions can be tough to visualize. Furthermore, the orientation of an object in space is important in physics research, for instance, spinors occupy an important place in quantum mechanics. Geometric algebra tools also provide the orientation of objects.

There are two expansions of this algebra: Conformal geometric algebra (CGA) and projective geometric algebra (PGA).

EPGA is based on duality: that is, we can represent work (wedge) done in one (exterior algebra) to be equivalent (join- $\vee$ ) in the other (dual exterior algebra). Similar to this situation, the meet $(\wedge)$ operator is also defined. The reason we work in dual space is because all of the Euclidean operations can be represented. Working in a projective dual space also prevents special cases from occurring.

In this paper, we use Euclidean curves to generate quaternion curves via geometric product. Thus, we describe what a quaternion operator looks like visually. Bearing this motivation, we wonder about the motion that will occur around a moving (dual) quaternion rather than around a fixed (dual) quaternion in the EPGA language.

## 2. Fundamentals

In this section, we try to explain the properties of GA and PGA. There are some operators. We explore how these operators lead to rotations just as complex numbers do.

One of the most remarkable works on quaternion algebra in the literature is Aslan and Yaylı, 22, where they define quaternion operators on curves and surfaces in Euclidean 3-space using geometric algebra. These operators generate motions that have orbits along the generated curve or surface and can be expressed as 1-parameter or 2-parameter homothetic motions. Besides, Shoemake presents a new kind of spline curve suitable for smoothly interpolating sequences of arbitrary rotations. The motion generated is smooth and natural, without quirks found in earlier methods, 23 .

Since geometric algebra is a very broad topic from kinematics $7-9$ to robot dynamics 10, from neuroscience 17 to modeling 11, our aim here is not to explain all the basic topics of geometric algebra. For more details, see $12,15,21$. Only definitions required in the article will be given.
2.1. Geometric algebra. Let $\mathbb{R}^{2}$, 2D-real space, be spanned by two linear independent orthonormal vectors: $\left\{\mathbf{e}_{1}, \mathbf{e}_{2}\right\}$. The inner and outer product of these elements produces new bases for geometric algebra:

$$
<\mathbf{e}_{1}, \mathbf{e}_{1}>=<\mathbf{e}_{2}, \mathbf{e}_{2}>=1 \text { and } \mathbf{e}_{1} \wedge \mathbf{e}_{2}
$$

Bearing with these elements we have the following bases for geometric algebra $\mathcal{G}(2,0)$ :

$$
\left\{1(\text { scalar }), \mathbf{e}_{1}, \mathbf{e}_{2}(\text { vectors }), \mathbf{e}_{1} \wedge \mathbf{e}_{2}(\text { pseudo-scalar })\right\}
$$

The algebra has also general elements called multivectors like that: $a 1+b \mathbf{e}_{1}+$ $c \mathbf{e}_{2}+d \mathbf{e}_{1} \wedge \mathbf{e}_{2}$, where $a, b, c, d \in \mathbb{R}$. Partially, $x 1+y \mathbf{e}_{1} \wedge \mathbf{e}_{2}=x+i y$, represent a complex number.
$\mathbb{R}^{3}, 3 D$-real space, is spanned by three independent vectors: $\left\{\mathbf{e}_{1}, \mathbf{e}_{2}, \mathbf{e}_{3}\right\}$. New basis for geometric algebra over $\mathbb{R}^{3}$ are:

$$
<\mathbf{e}_{1}, \mathbf{e}_{1}>=<\mathbf{e}_{2}, \mathbf{e}_{2}>=<\mathbf{e}_{3}, \mathbf{e}_{3}>=1
$$

and

$$
\mathbf{e}_{1} \wedge \mathbf{e}_{2}, \mathbf{e}_{1} \wedge \mathbf{e}_{3}, \mathbf{e}_{2} \wedge \mathbf{e}_{3}
$$

So vector spaces of the geometric algebra $\mathcal{G}(3,0)$ are

$$
\{\underbrace{1}_{\text {scalars }}, \underbrace{\mathbf{e}_{1}, \mathbf{e}_{2}, \mathbf{e}_{3}}_{\text {vector space }}, \underbrace{\mathbf{e}_{1} \wedge \mathbf{e}_{2}, \mathbf{e}_{2} \wedge \mathbf{e}_{3}, \mathbf{e}_{3} \wedge \mathbf{e}_{1}}_{\text {bivector space }}, \underbrace{\mathbf{e}_{1} \wedge \mathbf{e}_{2} \wedge \mathbf{e}_{3}}_{\text {trivector space }}\}
$$

All geometric algebra has $\sum_{k=0}^{3}\binom{3}{k}=2^{3}=8$ elements where the number of $k$-blades of geometric algebra over $\mathbb{R}^{3}$ is computed by $\binom{3}{k}$ combination. So any multi-vector is of the form: $a 1+b \mathbf{e}_{1}+c \mathbf{e}_{2}+d \mathbf{e}_{3}+e \mathbf{e}_{1} \wedge \mathbf{e}_{2}+f \mathbf{e}_{1} \wedge \mathbf{e}_{3}+g \mathbf{e}_{2} \wedge \mathbf{e}_{3}+$ $h \mathbf{e}_{1} \wedge \mathbf{e}_{2} \wedge \mathbf{e}_{3}$, where $a, b, c, d, e, f, g, h \in \mathbb{R}$.

Operations are important for the derivation of elements in a mathematical structure such as vector space or algebra. These processes should always show the feature of closure. The next definition gives the fundamental operator for geometric algebra.

Definition 1. Clifford defined the geometric product of two vectors, $\mathbf{u}$ and $\mathbf{v}$, as follows [9]: Let $\mathbf{u}=\sum_{i=1}^{n} u_{i} \mathbf{e}_{i}$ and $\mathbf{v}=\sum_{i=1}^{n} v_{i} \mathbf{e}_{i}$ then

$$
\begin{equation*}
\mathbf{u v}=<\mathbf{u}, \mathbf{v}>+\mathbf{u} \wedge \mathbf{v}=\sum_{i=1}^{n} u_{i} v_{i}+\sum_{i, j=1}^{n} u_{i} v_{j} \mathbf{e}_{i} \mathbf{e}_{j} \tag{1}
\end{equation*}
$$

where $\mathbf{e}_{i}$ are unit bases of $\mathbb{R}^{n}$. Here we use $\mathbf{e}_{i j}=\mathbf{e}_{i} \mathbf{e}_{j}=\mathbf{e}_{i} \wedge \mathbf{e}_{j}$ for simplicity.
Proposition 1. Let $\mathbf{u}, \mathbf{v}$ and $\mathbf{w}$ are vectors in $\mathbb{R}^{n}$ then following rules are provided
(i) Associativity:

$$
\mathbf{u}(\mathbf{v w})=\mathbf{u}(\mathbf{v}) \mathbf{w}=\mathbf{u} \mathbf{v} \mathbf{w}
$$

(ii) Distributivity:

$$
\mathbf{u}(\mathbf{v}+\mathbf{w})=\mathbf{u} \mathbf{v}+\mathbf{u w}
$$

and

$$
(\mathbf{v}+\mathbf{w}) \mathbf{u}=\mathbf{v} \mathbf{u}+\mathbf{w} \mathbf{u}
$$

(iii) Modulus:

$$
\|\mathbf{u}\|^{2}=\mathbf{u} \mathbf{u}=<\mathbf{u}, \mathbf{u}>
$$

Products of fundamental elements of geometric algebra are given as follows,

$$
\begin{gathered}
\mathbf{e}_{i} \mathbf{e}_{i}=1, \mathbf{e}_{i} \mathbf{e}_{j}=\mathbf{e}_{i} \wedge \mathbf{e}_{j}=-\mathbf{e}_{j} \wedge \mathbf{e}_{i}=-\mathbf{e}_{j} \mathbf{e}_{i} \\
\left(\mathbf{e}_{i} \wedge \mathbf{e}_{j}\right)\left(\mathbf{e}_{i} \wedge \mathbf{e}_{j}\right)=\mathbf{e}_{i} \mathbf{e}_{j} \mathbf{e}_{i} \mathbf{e}_{j}=-\mathbf{e}_{i} \mathbf{e}_{i}=-1 \\
\left(\mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3}\right)\left(\mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3}\right)=-\mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3} \mathbf{e}_{2} \mathbf{e}_{1} \mathbf{e}_{3}=-1
\end{gathered}
$$

where $i, j=1,2,3$ (for $i \neq j$ ).

Definition 2. In geometric algebra, there is a Hodge duality for elements and defined as follows [18]: Let $I=\mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3}$ be the pseudo-scalar then,

$$
1^{*}=-1 I=-I
$$

So, scalar and pseudo-scalar are Hodge dual of each other (Hodge duality generally represented by ${ }^{*}$ ). Similarly, vectors and blades are Hodge dual of each other:

$$
\begin{aligned}
& \mathbf{e}_{1}^{*}=-\mathbf{e}_{1} I=-\mathbf{e}_{1} \mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3}=-\mathbf{e}_{2} \mathbf{e}_{3}, \\
& \mathbf{e}_{2}^{*}=-\mathbf{e}_{2} I=-\mathbf{e}_{2} \mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3}=\mathbf{e}_{1} \mathbf{e}_{3}, \\
& \mathbf{e}_{3}^{*}=-\mathbf{e}_{3} I=-\mathbf{e}_{3} \mathbf{e}_{1} \mathbf{e}_{2} \mathbf{e}_{3}=-\mathbf{e}_{1} \mathbf{e}_{2} .
\end{aligned}
$$

Definition 3. Let $\mathbf{w}$ be a vector and $B=\mathbf{u} \wedge \mathbf{v}$ be a 2-vector, then

$$
\mathbf{w} B=\langle\mathbf{w}, B\rangle+\mathbf{w} \wedge B=(\mathbf{w} B)_{1}+(\mathbf{w} B)_{3}
$$

where $(.)_{k}$ represent the grade of element.
2.2. Projective geometric algebra. Although affine transformations of geometric objects can be achieved with vector algebra, this can cause some difficulties in 3D computer graphics, such as the algebraic separation of point and vector. For this, it is the algebraic structure that we call homogeneous coordinates and allows us to represent $n$-dimensional real space in ( $\mathrm{n}+1$ )-dimensional space. Let $\mathbf{x}=x_{1} \mathbf{e}_{1}+x_{2} \mathbf{e}_{2}+x_{3} \mathbf{e}_{3}$ be a point in Euclidean 3 -space then $X=\mathbf{x}+\mathbf{e}_{0}$ be a homogeneous point in 4D geometric algebra.

This space can also be defined in geometric algebra by adding an explicit extra basis: $\mathbf{e}_{0}$, satisfying $\mathbf{e}_{0}^{2}=0$, which corresponds to a null vector providing only linear terms expansion of an exponential function. So the metric structure would be $\mathbf{e}_{i} \cdot \mathbf{e}_{j}=\delta_{i j}$ for $i, j=1,2$ and $\mathbf{e}_{i} \cdot \mathbf{e}_{0}=0$. Since this metric structure is the same as the Euclidean metric, it also preserves isometry. There are also other bases like $\mathbf{e}^{2}=-1$ or 1 and generate higher dimensional projective geometric algebra. In general, these unusual bases are called geometric numbers. Summing up, its general notation in this point-based structure is $\mathbf{P}\left(\mathbb{R}_{p, n, z}\right)$ where $p, n, z$ stand for positive, negative and zero, respectively. Besides, plane-based model, for instance, the algebra $\mathbf{P}\left(\mathbb{R}_{2,0,1}^{*}\right)$ represents the proper 2 D Euclidean space. As far as we know from linear algebra, if $V$ is a vector space then there is a dual vector space $V^{*}$. So each geometric object in the exterior algebras in $\mathbf{P}(V)$ and $\mathbf{P}\left(V^{*}\right)$ has a representation in both. This is the Poincaré duality 13,14 .
$\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ provides coordinate-free, uniform representation for Euclidean elements: points, lines, and planes.

| $\wedge$ | $\mathbf{e}_{0}$ | $\mathbf{e}_{1}$ | $\mathbf{e}_{2}$ | $\mathbf{e}_{3}$ |
| :--- | :--- | :--- | :--- | :--- |
| $\mathbf{e}_{0}$ | 0 | $\mathbf{e}_{01}$ | $\mathbf{e}_{02}$ | $\mathbf{e}_{03}$ |
| $\mathbf{e}_{1}$ | $\mathbf{e}_{10}$ | 0 | $\mathbf{e}_{12}$ | $\mathbf{e}_{13}$ |
| $\mathbf{e}_{2}$ | $\mathbf{e}_{20}$ | $\mathbf{e}_{21}$ | 0 | $\mathbf{e}_{23}$ |
| $\mathbf{e}_{3}$ | $\mathbf{e}_{30}$ | $\mathbf{e}_{31}$ | $\mathbf{e}_{32}$ | 0 |

Although it has degenerate metric we can explain the reason why this algebra shows Euclidean isometries as follows: The basic elements in geometric algebra do not actually have Euclidean representations. Therefore, we can understand what they are geometrically by looking at their dual structure. However, dual PGA performs its operations directly with Euclidean elements. So we can call it EPGA for short. The basic linear elements of this algebra are planes (1-vector), and they are defined as follows,

$$
\mathbf{a}=\sum_{i=0}^{3} a_{i} \mathbf{e}_{i}
$$

It also includes meet and join operators. These operators decrease and increase of grades of elements of algebra, respectively. Thus, the union and intersection operations of points, lines, and planes in PGA can be done with wedge and progressive product, respectively.

Quaternions are also even subalgebra (zero and two graded) of projective geometric algebra $\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ and a quaternion is represented by

$$
q=a+b \mathbf{e}_{12}+c \mathbf{e}_{31}+d \mathbf{e}_{23}
$$

where $a, b, c, d \in \mathbb{R}$.

## 3. Quaternion Curves and Surfaces

Grounded in the elegant framework of geometric algebra, quaternion curves and surfaces stand as a cornerstone in the field of 3D geometry and computer graphics. In this chapter, we enter a world where advanced mathematics meets practical applications, exploring the profound implications of quaternions in representing and manipulating curves and surfaces.

As a hyper-complex extension of complex numbers, quaternions offer a concise and efficient way to handle 3D rotations and orientations, finding extensive applications in fields ranging from computer graphics and robotics to physics simulations. Represented by geometric algebra, these quaternions are not just abstract mathematical constructs, but powerful tools that allow us to describe the complex motion of objects in space with remarkable precision and flexibility.

Definition 4. Let $\mathbf{a}(t), \mathbf{b}(t)$ be vector fields then

$$
q(t)=\mathbf{a}(t) \mathbf{b}(t)=<\mathbf{a}(t), \mathbf{b}(t)>+\mathbf{a}(t) \wedge \mathbf{b}(t)
$$

is a quaternion curve and its conjugate is given by reverse order product

$$
\widetilde{q}(t)=\mathbf{b}(t) \mathbf{a}(t)=<\mathbf{b}(t), \mathbf{a}(t)>+\mathbf{b}(t) \wedge \mathbf{a}(t)
$$

Definition 5. Let $\mathbf{a}(t), \mathbf{b}(s)$ be vector fields then

$$
q(t, s)=\mathbf{a}(t) \mathbf{b}(s)=<\mathbf{a}(t), \mathbf{b}(s)>+\mathbf{a}(t) \wedge \mathbf{b}(s)
$$

is a quaternion surface.
To see the behavior of these quaternion curves in 3-dimensional space, it is necessary to apply them to a point. Let us now formulate here the rotation of a point through a quaternion operator generated by the meeting of two unit vectors in space:

Let $\|\mathbf{u}\|=\|\mathbf{v}\|=1, q=\mathbf{u v}=\cos (\theta)+\sin (\theta) \mathrm{B}$ and its conjugate $\widetilde{q}=\mathbf{v u}=$ $\cos (\theta)-\sin (\theta)$ B. Take a point in EPGA as $P=\mathbf{e}_{123}+P_{E}$ then the rotated point $P_{r}$ is as follows;

$$
\begin{equation*}
P_{r}=q P \widetilde{q}=\mathbf{e}_{123}+q P_{E} \widetilde{q} \tag{2}
\end{equation*}
$$

Thus, applying real quaternion operators to EPGA elements gives the same result as in EGA.
3.1. Trajectory curves. The concept of trajectories generated by quaternion curves involves representing rotations in 3D space using quaternions and understanding how these rotations affect the orientation of objects over time.

In the context of trajectories, quaternions are used to smoothly interpolate between different orientations of an object, creating a continuous curve that describes the object's rotation over time. Quaternions have certain advantages over other rotation representations (such as Euler angles) because they do not suffer from gimbal lock and provide smooth interpolation without singularities.
Corollary 1. Let $q(t)$ be a quaternion curve and $P$ be a point in $\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ then

$$
\begin{equation*}
\alpha(t)=q(t) P \widetilde{q}(t) \tag{3}
\end{equation*}
$$

is a trajectory curve.
Example 1. Let $\mathbf{u}(t)=\cos (t) \mathbf{e}_{1}+\sin (t) \mathbf{e}_{2}, \mathbf{v}(t)=\cos (t) \mathbf{e}_{1}+\sin (t) \mathbf{e}_{3}, P=\mathbf{e}_{123}+$ $\mathbf{e}_{032}$ be unit vector fields and a point in $\in \mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$, respectively. Then

$$
\begin{aligned}
q(t) & =\mathbf{u}(t) \mathbf{v}(t) \\
& =<\cos (t) \mathbf{e}_{1}+\sin (t) \mathbf{e}_{2}, \cos (t) \mathbf{e}_{1}+\sin (t) \mathbf{e}_{3}> \\
& +\left(\cos (t) \mathbf{e}_{1}+\sin (t) \mathbf{e}_{2}\right) \wedge \cos (t) \mathbf{e}_{1}+\sin (t) \mathbf{e}_{3} \\
& =\cos (t)^{2}+\cos (t) \sin (t) \mathbf{e}_{1} \mathbf{e}_{3}-\cos (t) \sin (t) \mathbf{e}_{1} \mathbf{e}_{2} \\
& +\sin ^{2}(t) \mathbf{e}_{2} \mathbf{e}_{3} .
\end{aligned}
$$



Figure 1. The curve that is generated by the quaternion curve.

So the curve of quaternion rotation is (see Fig.1)

$$
\begin{aligned}
\alpha(t) & =q(t) P \widetilde{q}(t) \\
& =\mathbf{e}_{123}+\cos (2 t)^{2} \mathbf{e}_{230}+\frac{\sin (4 t)}{2} \mathbf{e}_{301}-\sin (2 t) \mathbf{e}_{012}
\end{aligned}
$$

This curve is also called Viviani's curve.
3.2. Trajectory surfaces. Similar situations to the trajectory curves can also be done for surfaces. The only difference here is that the parameters of the vector fields forming the quaternion are different from each other.

Corollary 2. Let $q(t, s)$ be a quaternion surface and $P$ be a point in $\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ then

$$
\begin{equation*}
X(t, s)=q(t, s) P \widetilde{q}(t, s) \tag{4}
\end{equation*}
$$

is a trajectory surface.
Corollary 3. Let $q_{\theta}$ be a quaternion, where $\theta$ is the angle between vectors that are constructing the quaternion, and $\alpha(t)$ be a regular curve in $\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ then

$$
\begin{equation*}
\beta(t, \theta)=q_{\theta} \alpha(t) \widetilde{q_{\theta}} \tag{5}
\end{equation*}
$$

is a rotational surface. This is also a special case of trajectory curves.

Corollary 4. Let $q(t)$ be a quaternion curve $\alpha(s)$ be a regular curve in $\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ then

$$
\begin{equation*}
\beta(t, s)=q(t) \alpha(s) \widetilde{q}(t) \tag{6}
\end{equation*}
$$

is a trajectory surface.
Let's define some surfaces as trajectory surfaces with the tool we have generated.
3.2.1. Sphere as a trajectory surface. This example can be used for surfaces (one can take a different parameter for the second vector field as we defined in Def 3.2). The resulting trajectory surface is this time, a 2 -sphere. This is just like the product of two curves in the topological sense: $\mathbb{S}^{2} \simeq \mathbb{S}^{1} \times \mathbb{S}^{1}$.
3.2.2. Cone as a trajectory surface. It is easy to obtain a cone surface using this tool. As is well known, a cone is formed by a line passing through two points in space, one of the points being fixed and the other orbiting on a circle. We can make this imaginary geometric idea as it is with the help of geometric algebra. The flexibility in the choice of the vector fields that make up the quaternion and the object to be rotated in this tool offers us different ways and offers us the opportunity to create the geometric object to be created in different ways. Everything depends on your imagination. For example, for this cone surface, we can choose one of the vector fields forming the quaternion as a unit circle and the other as a point perpendicular to the plane in which this circle is located. The geometric object to be rotated is then a line.
3.3. Dual quaternions and rigid motions. The projective 4D analog of a quaternion is called a dual quaternion. This is where the real difference of EPGA comes into play. So all Euclidean motions can be described in this space and we can assume the dual number unit $\epsilon$ as a pseudo-scalar $\mathbf{e}_{0123}$, i.e. provide the mystique properties of its: $\mathbf{e}_{0123}^{2}=0$. Thus, we define the dual element as an algebraic basis.

Let $q=x_{0}+x_{1} e_{2} e_{3}+x_{2} e_{3} e_{1}+x_{3} e_{1} e_{2}, r=y_{0}+y_{1} e_{2} e_{3}+y_{2} e_{3} e_{1}+y_{3} e_{1} e_{2}$ be two quaternions. Then we can construct the dual quaternions in geometric algebra way as follows 10:

$$
\begin{aligned}
Q & =q+r \mathbf{e}_{0123}=q-r^{*} \\
& =x_{0}+x_{1} e_{2} e_{3}+x_{2} e_{3} e_{1}+x_{3} e_{1} e_{2} \\
& +y_{0} e_{0} e_{1} e_{2} e_{3}+y_{1} e_{1} e_{0}+y_{2} e_{2} e_{0}+y_{3} e_{3} e_{0}
\end{aligned}
$$

Therefore, $P_{m}=Q P \bar{Q}$ represents a rigid transformation of a point in the space.
Definition 6. Let $q(t), r(t)$ be quaternion curves, then

$$
Q(t)=q(t)+r(t) \mathbf{e}_{0123}
$$

is a dual quaternion curve.


Figure 2. Trajectory curves that are generated by the dual quaternion.

Corollary 5. Let $Q(t)$ be dual quaternion curve and $P$ be a point in $\mathbf{P}\left(\mathbb{R}_{3,0,1}^{*}\right)$ then

$$
X(t)=Q(t) P \bar{Q}(t)
$$

is a trajectory curve.
Example 2. For the most basic situation, let $Q(t)=\cos (t) \mathbf{e}_{12}+\sin (t) \mathbf{e}_{23}+\mathbf{e}_{01}+$ $\mathbf{e}_{0123}$ be a dual quaternion curve, then for points $P_{x}=\mathbf{e}_{123}+\mathbf{e}_{032}, P_{y}=\mathbf{e}_{123}+\mathbf{e}_{013}$, $P_{z}=\mathbf{e}_{123}+\mathbf{e}_{102}$ there are three trajectory curves, see Fig.2.

An argument similar to the one above can be constructed with the basis (1vectors) of EPGA. At this time, one can get a line that does not pass through the origin by meeting the two 1-vectors: Let $\mathbf{a}, \mathbf{b}$ be two 1 -vectors and at least one of them is not passing through the origin, then their geometric product is

$$
\mathbf{a b}=q_{0}+\mathcal{L}
$$

where $q_{0}$ and $\mathcal{L}$ are scalar and plücker coordinates of the line, respectively. This operator generates a trajectory (orbit) of a point around a moving line in space.

It is decided whether a quaternion operator formed by the geometric multiplication of two 1 -vectors is real or dual, by looking at whether its first terms (zeroth index) are zero. In other words, the line obtained from the intersection (i.e., the meet operator) of two planes passing through the origin represents the quaternion operator that represents a rotation around this line passing through the origin. On
the other hand, if one of the first terms is nonzero, the resulting quaternion will represent a screw motion. Therefore, this quaternion acts like a dual quaternion operator.

## 4. Conclusion

Geometric algebra has lately garnered significant attention due to its profound applications for both imaging and performing fine operations on geometric objects. The field has seen remarkable progress, particularly in enhancing our capability to fantasize about complex geometric generalities. In this study, we embarked on a disquisition that extended the operation of geometric algebra from traditional vectorground representations of curves and surfaces in classical figures to the realm of quaternions. Our purpose is to demonstrate the unique capabilities of quaternions as an important fine driver, shedding light on their part in generating topological structures when applied to classical 3D geometric objects.

The implications of our findings extend far beyond the realm of classical Euclidean geometry. While our study primarily focused on classical 3D geometric objects, the inherent flexibility of quaternions suggests that similar investigations can be carried out in the domain of non-Euclidean geometries. This exciting prospect hints at a wealth of fascinating results waiting to be uncovered, as the interplay between quaternions and non-Euclidean geometries promises to yield profound insights and applications in various scientific and engineering disciplines.

In summary, our research represents a pivotal contribution to the field of geometric algebra by showcasing the remarkable utility of quaternions as operators in transforming classical geometric objects and elucidating the emergence of topological structures. This work not only deepens our understanding of the relationship between algebra and geometry but also opens up a tantalizing avenue for future research, where quaternions can be harnessed to explore the rich landscapes of nonEuclidean geometries, potentially revolutionizing how we perceive and interact with the mathematical underpinnings of the physical world.

Declaration of Competing Interests The author has no competing interests to declare.

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# EXPONENTIATED GENERALIZED RAMOS-LOUZADA DISTRIBUTION WITH PROPERTIES AND APPLICATIONS 

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#### Abstract

In this paper, we propose a new generalization of Ramos-Louzada (RL) distribution based on two additional shape parameters. Along with the genesis of its distributional form, the derivation of cumulative density function (cdf), survival and hazard rate functions, the quantile function (qf), moments, moment generating function (mgf), Shannon and Renyi entropies, order statistics and a linear representation of the proposed distribution are inspected. Several estimation methods of the model parameters are discussed throughout two comprehensive simulation studies conducted to compare its performance against some lifetime distributions. Application of a real dataset is presented to illustrate the potentiality of this distribution in line with the simulation studies.


## 1. Introduction

Lifetime modeling of complex studies has created a growing interest in the generation of flexible distributions that can provide solutions to certain problems of lifetime systems. Ramos-Louzada is such a distribution recently proposed by Ramos and Louzada ( $\mid 24)$ to take instantaneous failures into account that can inevitably occur in many lifetime applications. It is announced to be a worthwhile alternative to the Exponential and Lindley ( 19 ) distributions and take the forms of both with a shape parameter $\lambda \geq 2$. That is, the distribution becomes the exponential distribution for large values of $\lambda$ and it resembles to the Lindley distribution as $\lambda$ decreases towards 2.

[^5]Let the random variable $X$ follows the RL distribution with the rate parameter $\theta=\frac{1}{\lambda}, \lambda \geq 2 ;$

$$
\begin{equation*}
g(x)=\frac{\left(\theta^{2} x-2 \theta+1\right) \theta}{1-\theta} e^{-\theta x} \tag{1}
\end{equation*}
$$

where $x \geq 0$ and $0.5 \geq \theta>0$. The cdf of $X \sim R L(\theta)$ is defined as

$$
\begin{equation*}
G(x)=1-\frac{\left(\theta^{2} x-\theta+1\right)}{1-\theta} e^{-\theta x} \tag{2}
\end{equation*}
$$

Although the RL distribution is attractive for its simplicity, it fails to provide precise evaluation of many lifetime datasets, since it contains only one parameter. Many researchers benefit from generalizing baseline (stated otherwise parent or target) distributions by adding one or more parameters into the model to increase the model fit and overcome the absence of sufficient flexibility in modeling the data. In this respect, Al-Mofleh et al. ( 3 ) recently proposed a two-parameter generalization of the RL distribution by inserting a power parameter into the model. They showed that the generalized Ramos-Louzada (GRL) distribution performs better than some well-known distributions such as Marshall-Olkin ( 21 ), exponentiated exponential ( 11 ) and generalized Lindley ( 23 ) distributions with respect to some bias and accuracy measures.

This paper proposes a new three-parameter model as a competetive extension for this generalization of the RL distribution, namely the exponentiated generalized Ramos-Louzada (EGRL) distribution. The new distribution relies on the class of distributions established by Corderio et al. ( 7 ). The usual definition of the probability density function (pdf) of this family of distributions is

$$
\begin{equation*}
f(x)=\alpha \beta g(x)[1-G(x)]^{\alpha-1}\left(1-[1-G(x)]^{\alpha}\right)^{\beta-1} \tag{3}
\end{equation*}
$$

where $\alpha>0$ and $\beta>0$ are two shape parameters and $g(x)$ and $G(x)$ are the pdf and cdf of the baseline distribution, respectively. The shape parameters $\alpha$ and $\beta$ in equation (3) provide better flexibility in the tails of the data and increase the entropy in the center ( $7, \mathrm{p} .2]$ ). The cdf of the family of distributions is of the form

$$
\begin{equation*}
F(x)=\left(1-[1-G(x)]^{\alpha}\right)^{\beta} \tag{4}
\end{equation*}
$$

Our basic motivation for such generalization is to provide a better fit of RL distribution to the wider range of problems in statistics. It is also of our goal to achieve reliable estimation of model parameters considering various estimation methods. This is particularly important as it affects the model selection process. Evaluation of model fit via goodness of fit statistics is a usual practice in the literature. Al-Mofleh et al. ( 3 ) consider only the minus log likelihood ( $-\ell$ ), Cramer-von Mises (C*; 9) and Kolmogorov-Smirnov (KS*; 16, 31) goodness of
fit statistics. The assessment of model fit via these goodness of fit statistics might produce biased results, since they do not take the model complexity into account when choosing the best distribution in a set of distributions. The distributions under consideration should also be compared to each other by means of using information criteria.

Incorporating additional adequate parameter(s) into the model improves the model fit and provides more flexibility in analyzing datasets. However, caution should be taken when generalizing baseline distributions using more parameters in the model. Achieving a good model fit requires taking into account the balance between the sample size and the number of parameters in the model (bias versus variance tradeoff as used in the literature). The information criteria such as Akaike information criterion (AIC; 1, 22) and Bayesian information criterion (BIC; 29) are originally developed for solving this problem as they do not only rely on log likelihood values, but also on penalty values. The log likelihood represents the fit of a model to the data at hand as the penalty value penalizes the model depending on (a function of) the number of parameters in the model. Thus, we compare the EGRL distribution against a set of alternative distributions by means of not only using model fit statistics, but also different types of information criteria.

The outline of the paper is as follows. In Section 2, we derive various statistical and reliability properties of the EGRL distribution. In Section 3, we elaborate on the methods of maximum likelihood estimation (MLE), least squares estimation (LSE), weighted least squares estimation (WLSE), and Cramer-von Mises estimation (CVME) to obtain the estimates of model parameters and their standard errors for the EGRL distribution. In Section 4, we perform two simulation studies. In the first simulation study, we evaluate the performance of the methods in estimating the parameters of the EGRL distribution with respect to bias, precision, and accuracy measures. In the second simulation study, we compare the performance of the EGRL distribution to that of a set of other lifetime distributions with respect to some goodness of fit statistics and information criteria for each estimation method. In Section 5, we exemplify the applicability of EGRL distribution for a real life problem. We illustrate that the goodness of fit statistics may not be able to detect the best distribution in a set of distributions and information criteria should be used instead when comparing the performance of distributions. The paper will be concluded with a short discussion.

## 2. The EGRL Distribution

2.1. Probability density and cumulative density functions. Incorporating equations (1) and (2) into the general definition in equation (3), we obtain the pdf of EGRL distribution which is given by

$$
f(x)=\alpha \beta \frac{\left(\theta^{2} x-2 \theta+1\right) \theta}{1-\theta} e^{-\theta x}\left[\frac{\left(\theta^{2} x-\theta+1\right)}{1-\theta} e^{-\theta x}\right]^{\alpha-1}
$$

$$
\begin{equation*}
\times\left(1-\left[\frac{\left(\theta^{2} x-\theta+1\right)}{1-\theta} e^{-\theta x}\right]^{\alpha}\right)^{\beta-1} \tag{5}
\end{equation*}
$$

For $\alpha=\beta=1$, the distribution reduces to the RL distribution. Similarly, by replacing $G(x)$ in equation (4) with the cdf of RL distribution in equation (2), we obtain the cdf of EGRL distribution as

$$
\begin{equation*}
F(x)=\left(1-\left[\frac{\left(\theta^{2} x-\theta+1\right) e^{-\theta x}}{1-\theta}\right]^{\alpha}\right)^{\beta} \tag{6}
\end{equation*}
$$



Figure 1. The pdf and cdf plots of EGRL distribution with varying values of $\alpha, \beta$ and $\theta$ parameters.

Figure 1 displays the plots for the pdf and cdf of EGRL distribution using different values of $\alpha, \beta$ and $\theta$ parameters. As can be seen on the left panel of the figure, the EGRL distribution is flexible in the sense that it can be positively skewed with or without reversed-J shape. The plots on the right panel of the figure show that the cdf of EGRL distribution increases towards one with increasing values of the random variable $X$ for varying values of parameters $\alpha, \beta$ and $\theta$.
2.2. Survival and hazard rate functions. The survival function (stated otherwise reliability function) is commonly used for lifetime datasets which often represents the probability of a patient's survival or an object's resistance until a predetermined time point. The survival function of EGRL distribution indicating the complement of the cdf in equation (6) is given by

$$
\begin{equation*}
S(x)=1-\left(1-\left[\frac{\left(\theta^{2} x-\theta+1\right) e^{-\theta x}}{1-\theta}\right]^{\alpha}\right)^{\beta} \tag{7}
\end{equation*}
$$



Figure 2. The survival and hazard rate plots of EGRL distribution with varying values of $\alpha, \beta$ and $\theta$ parameters.

Another widely used tool that can serve to characterize the EGRL distribution is the hazard rate function which indicates the probability of the occurence of an event. The values of hazard rate function for the EGRL distribution can easily be obtained by

$$
\begin{equation*}
H(x)=\frac{f(x)}{S(x)}, \tag{8}
\end{equation*}
$$

where $f(x)$ is the pdf in equation (5) and $S(x)$ is the survival function in equation (7).

Figure 2 displays the plots of survival and hazard rate functions for the EGRL distribution. These plots exhibit increasing, decreasing, and reversed-J shaped hazard rate functions and decreasing survival functions with increasing values of random variable X .
2.3. The quantile function. The quantile function (qf) of EGRL distribution is the inverse of the cdf in equation (6). By applying $w=-\theta x-\frac{1-\theta}{\theta}$ transformation and using $w$ in Lambert form $w e^{w}$ for $u=G(x)$ in equation (2), we obtain

$$
\begin{equation*}
w e^{w}=\frac{(1-\theta)(u-1) e^{1-\frac{1}{\theta}}}{\theta} \tag{9}
\end{equation*}
$$

This means that $w$ can be defined as a Lambert function of the real argument $w e^{w}$. The real argument $w e^{w} \in\left(-\frac{1}{e}, 0\right)$ for $u \in(0,1)$. Thus,

$$
\begin{equation*}
Q_{E G R L}(u)=\frac{-\theta W_{-1}\left[\frac{(\theta-1)\left[\left(1-u^{\frac{1}{\beta}}\right)^{\frac{1}{\alpha}}\right] e^{1-\frac{1}{\theta}}}{\theta}\right]+\theta-1}{\theta^{2}} \tag{10}
\end{equation*}
$$

where $W_{-1}$ is the negative branch of the Lambert function, $0<\theta \leq 0.5$, and $0<u<1$ (see [7, p. 2]). The values of the negative branch of Lambert function $W_{-1}$ can easily be obtained using lambertWm1 subroutine of lamW package in $R$ statistical software.

The Bowley skewness ( 15 ) and Moorsis kurtosis ( 22 ) measures for EGRL distribution are defined by

$$
\begin{equation*}
B=\frac{Q_{E G R L}(3 / 4)+Q_{E G R L}(1 / 4)-2 Q_{E G R L}(2 / 4)}{Q_{E G R L}(3 / 4)-Q_{E G R L}(1 / 4)} \tag{11}
\end{equation*}
$$

and

$$
\begin{equation*}
M=\frac{Q_{E G R L}(3 / 8)-Q_{E G R L}(1 / 8)+Q_{E G R L}(7 / 8)-Q_{E G R L}(5 / 8)}{Q_{E G R L}(6 / 8)-Q_{E G R L}(2 / 8)}, \tag{12}
\end{equation*}
$$

where, for example, $Q_{E G R L}(3 / 4)$ is the third quartile and $Q_{E G R L}(5 / 8)$ is the fifth octile of the qf for the EGRL distribution. Table 1 shows how these measures behave with varying values of parameters $\alpha, \beta$ and $\theta$. In line with the pdf plots in Figure 1, increasing values of Moorsis kurtosis measure are associated with the pdfs with heavier tails. Positive values of the Bowley skewness measure in this table indicate that the distributions are right skewed.

Table 1. The Bowley skewness and Moorsis kurtosis measures with varying values of $\alpha, \beta$, and $\theta$ parameters.

| $\alpha$ | $\beta$ | $\theta$ | Bowley skewness | Moorsis kurtosis |
| :---: | :---: | :---: | :---: | :---: |
| 0.1 | 1.0 | 0.5 | 0.23 | 1.25 |
| 0.5 | 1.0 | 0.5 | 0.19 | 1.30 |
| 1.0 | 1.0 | 0.5 | 0.17 | 1.32 |
| 1.5 | 1.0 | 0.5 | 0.15 | 1.33 |
| 2.0 | 1.0 | 0.5 | 0.15 | 1.34 |
| 1.0 | 0.1 | 0.5 | 0.77 | 2.49 |
| 1.0 | 0.5 | 0.5 | 0.24 | 1.16 |
| 1.0 | 1.5 | 0.5 | 0.14 | 1.46 |
| 1.0 | 2.0 | 0.5 | 0.13 | 1.56 |
| 2.0 | 2.0 | 0.1 | 0.18 | 1.31 |
| 2.0 | 2.0 | 0.2 | 0.17 | 1.32 |
| 2.0 | 2.0 | 0.3 | 0.15 | 1.33 |
| 2.0 | 2.0 | 0.4 | 0.13 | 1.41 |
| 2.0 | 2.0 | 0.5 | 0.12 | 1.61 |

2.4. Moments. We follow an analogous procedure to the one given in the previous subsection with a slightly different transformation

$$
\begin{equation*}
v=1-G(x)=\frac{\left(\theta^{2} x-\theta+1\right)}{1-\theta} e^{-\theta x} \tag{13}
\end{equation*}
$$

where $0<v<1$. The $m$ th moment of EGRL distribution is given by

$$
\begin{equation*}
E\left(X^{m}\right)=\int_{0}^{\infty} x^{m} f(x) d x=\alpha \beta \int_{0}^{1}(-1)^{m+1}[z(v)]^{m} v^{\alpha-1}\left(1-v^{\alpha}\right)^{\beta-1} d v \tag{14}
\end{equation*}
$$

where

$$
z(v)=\frac{-\theta W_{-1}\left[\frac{(1-\theta) v e^{1-\frac{1}{\theta}}}{\theta}\right]+\theta-1}{\theta^{2}} .
$$

The next subsection recalls some useful definitions and power series expansions that can be used to obtain the moments, mgf, Shannon and Renyi entropies, and order statistics of EGRL distribution.
2.5. Useful definitions and power series expansions. Let T be a random variable from the exponentiated exponential distribution which has the following pdf

$$
\begin{equation*}
r(t)=\alpha \beta e^{-\alpha t}\left(1-e^{-\alpha t}\right)^{\beta-1} \tag{15}
\end{equation*}
$$

and cdf

$$
\begin{equation*}
R(t)=\left(1-e^{-\alpha t}\right)^{\beta} \tag{16}
\end{equation*}
$$

where $\alpha, \beta, t>0(\mid 11)$. The pdf of a random variable from the exponentiated generalized family of distributions can also be defined as

$$
\begin{equation*}
f(x)=\frac{g(x)}{1-G(x)} r(-\log [1-G(x)]) \tag{17}
\end{equation*}
$$

where $T=-\log [1-G(X)]$ follows the exponentiated exponential distribution ( 4 ). By using equation (17), $u=G(x)$, and the power series expansion

$$
\begin{equation*}
-\log (1-u)=\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1} \tag{18}
\end{equation*}
$$

the pdf $r(-\log [1-G(x)])$ becomes

$$
\begin{equation*}
r(-\log [1-G(x)])=r\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right)=\alpha \beta e^{-\alpha D}\left(1-e^{-\alpha D}\right)^{\beta-1} \tag{19}
\end{equation*}
$$

where $D=\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}$. Similarly, by using equation $18, u=G(x)$, and the power series expansion above, the corresponding cdf is defined as

$$
\begin{equation*}
R(-\log [1-G(x)])=R\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right)=\left(1-e^{-\alpha D}\right)^{\beta} \tag{20}
\end{equation*}
$$

By applying another useful power series expansion

$$
(1-y)^{a}=\sum_{k=0}^{\infty}\binom{a}{k}(-1)^{k}|y|,|y|<1
$$

we obtain

$$
\begin{equation*}
[1-F(x)]^{n-r}=\sum_{j=0}^{n-r}\binom{n-r}{j}(-1)^{j} F^{j}(x) \tag{21}
\end{equation*}
$$

which in turn will be used particularly to establish the order statistics for the EGRL distribution.
2.6. Moments using power series expansions and quantile function. By using equations (19) and (21), the $m$ th moment of EGRL distribution can also be defined as follows:

$$
\begin{align*}
E\left(X^{m}\right) & =\int_{0}^{\infty} x^{m} \frac{g(x)}{1-G(x)} r(-\log [1-G(x)]) d x=\int_{0}^{1} Q^{m}(u) \frac{1}{1-u} r\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right) d u \\
& =\alpha \beta \int_{0}^{1} Q^{m}(u) \frac{1}{1-u} e^{-\alpha D}\left(1-e^{-\alpha D}\right)^{\beta-1} d u \tag{22}
\end{align*}
$$

where $D=\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}$. Here, $Q(u)=G^{-1}(u)=x$ is the quantile function (i.e., the inverse of the cdf) in equation (9), so that $u=G(x)$ and $d u=g(x) d x$.
2.7. Moment generating function. Following the procedure used to obtain the moments in the previous subsection, the mgf of EGRL distribution can be obtained as

$$
\begin{align*}
E\left(e^{b X}\right) & =\int_{0}^{\infty} e^{b x} \frac{g(x)}{1-G(x)} r(-\log [1-G(x)]) d x=\int_{0}^{1} e^{b Q(u)} \frac{1}{1-u} r\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right) d u \\
& =\alpha \beta \int_{0}^{1} e^{b Q(u)} \frac{1}{1-u} e^{-\alpha D}\left(1-e^{-\alpha D}\right)^{\beta-1} d u \tag{23}
\end{align*}
$$

2.8. Shannon entropy. The Shannon entropy ( 30 ) is a measure to ascertain the information provided by a random variable. The Shannon entropy for the random variable X from the EGRL distribution is given by

$$
\begin{equation*}
\eta_{S}=-E\left(\log \left[\frac{g(x) r(t)}{1-G(x)}\right]\right) \tag{24}
\end{equation*}
$$

where $r(t)$ is used as a generator to attain the family of distributions in equation (3).

The association between the Shannon entropy for the generator variable T which has the support $[0, \infty]$ and the variable $X$ from the beta-exponential-X family of distributions can be defined by using $T=-\log [1-G(X)]$, and thus, $X=G^{-1}\left(1-e^{-T}\right)$ ( 4 ). This association also applies to our case for which the variable $T$ is from the exponentiated exponential distribution which has the support $[0, \infty]$ and the variable $X$ is from the EGRL distribution, since the exponentiated generalized family
of distributions is a special case of the beta-exponential-X family of distributions. Thus, the Shannon entropy above is defined as

$$
\begin{align*}
\eta_{S} & =-E\left(\log f\left[G^{-1}\left(1-e^{-T}\right)\right]\right)+\eta_{T}-\mu_{T} \\
& =-E\left(\log f\left[G^{-1}\left(1-e^{-T}\right)\right]\right) \\
& +\log \left[(\alpha \beta)^{-1}\right]+\beta \Psi(\beta+1)-(\beta-1) \Psi(\beta)-\Psi(1) \\
& +\frac{\Psi(\beta+1)-\Psi(1)}{\alpha} \tag{25}
\end{align*}
$$

where $\eta_{T}=\log \left[(\alpha \beta)^{-1}\right]+\beta \Psi(\beta+1)-(\beta-1) \Psi(\beta)-\Psi(1)$ is the Shannon entropy for random variable $\mathrm{T}, \mu_{T}=\frac{\Psi(\beta+1)-\Psi(1)}{\alpha}$ is its mean and $\Psi($.$) is the digamma$ function [4, p. 68].
2.9. Renyi entropy. The Renyi entropy ( 26 ) is another widely used measure to quantify the information in random variables. The Renyi entropy is an extension of the Shannon entropy. The Renyi entropy of order $\gamma$ for the random variable $X$ from the EGRL distribution is given by

$$
\begin{align*}
\eta_{R} & =\frac{1}{1-\gamma} \log \int_{0}^{\infty} f^{\gamma}(x) d x=\frac{1}{1-\gamma} \log \int_{0}^{\infty} \frac{g^{\gamma}(x)}{1-G^{\gamma}(x)} r^{\gamma}(-\log [1-G(x)]) d x \\
& =\frac{1}{1-\gamma} \log \int_{0}^{1} \frac{g^{\gamma-1}[Q(u)]}{1-u^{\gamma}} r^{\gamma}\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right) d u \\
& =\frac{\alpha^{\gamma} \beta^{\gamma}}{1-\gamma} \log \int_{0}^{1} \frac{g^{\gamma-1}[Q(u)]}{1-u^{\gamma}} e^{-\alpha \gamma D}\left(1-e^{-\alpha D}\right)^{\gamma(\beta-1)} d u \tag{26}
\end{align*}
$$

where $g(x)$ is the pdf of RL distribution and $D=\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}$. The value of order $\gamma$ influences the information obtained from random variable $X$. The Renyi entropy recovers the minimum entropy if $\gamma=\infty$, the maximum entropy if $\gamma=0$, and Shannon's entropy if $\gamma \rightarrow 1$ ( 27$)$.
2.10. Order statistics. Let $X_{(1)}=\min \left\{X_{1}, X_{2}, \ldots, X_{n}\right\}$ and $X_{(n)}=\max \left\{X_{1}, X_{2}\right.$, $\left.\ldots, X_{n}\right\}$ are the smallest and largest values of a random sample $X_{1}, X_{2}, \ldots, X_{n}$, respectively. In line with Arnold et al. ( $\sqrt{5}$ ), the pdf of the $r$ th order statistic (i.e, the $r$ th smallest value) is defined by

$$
\begin{equation*}
f_{(r)}(x)=\binom{n}{r} F^{r-1}(x)[1-F(x)]^{n-r} f(x) \tag{27}
\end{equation*}
$$

By applying the power series expansion in equation (23), the pdf of the $r$ th order statistic is defined by

$$
f_{(r)}(x)=\binom{n}{r} F^{r-1}(x)[1-F(x)]^{n-r} f(x)
$$

$$
\begin{align*}
& =\frac{n!}{(r-1)!(n-r)!} \sum_{j=0}^{n-r}\binom{n-r}{j}(-1)^{j} F^{j+r-1}(x) f(x) \\
& =\frac{n!}{(r-1)!(n-r)!} \sum_{j=0}^{n-r}\binom{n-r}{j}(-1)^{j} R^{j+r-1}(-\log [1-G(x)]) \\
& \times \frac{g(x)}{1-G(x)} r(-\log [1-G(x)]) \\
& =\frac{n!}{(r-1)!(n-r)!} \sum_{j=0}^{n-r}\binom{n-r}{j}(-1)^{j} R^{j+r-1}\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right) \\
& \times \frac{g(x)}{1-G(x)} r\left(\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}\right) \\
& =\frac{n!}{(r-1)!(n-r)!} \sum_{j=0}^{n-r}\binom{n-r}{j}(-1)^{j}\left(1-e^{-D}\right)^{\beta(j+r-1)} \\
& \times \frac{g(x)}{1-G(x)} e^{-\alpha D}\left(1-e^{-\alpha D}\right)^{\beta-1}, \tag{28}
\end{align*}
$$

where $g(x)$ and $G(x)$ are the pdf and cdf of RL distribution and $D=\sum_{i=0}^{\infty} \frac{u^{i+1}}{i+1}$.
2.11. Linear representation. Corderio and Lemonte ( 8 ) produce the linear representations of equations (3) and (4). We summarize their procedure here by using $G(x)$ as the cdf of the baseline RL distribution. By applying the generalized binomial expansion in equation (23) twice in equation (4), the cdf of EGRL distribution is defined as

$$
\begin{aligned}
F(x) & =\left(1-[1-G(x)]^{\alpha}\right)^{\beta}=\sum_{k=0}^{\infty}(-1)^{k}\binom{\beta}{k}[1-G(x)]^{\alpha k} \\
& =\sum_{k=0}^{\infty}(-1)^{k}\binom{\beta}{k} \sum_{j=0}^{\infty}(-1)^{j}\binom{\alpha k}{j} G^{j}(x)=\sum_{j=0}^{\infty} \sum_{k=1}^{\infty}(-1)^{j+k+1}\binom{\beta}{k}\binom{\alpha k}{j+1} G^{j+1}(x),
\end{aligned}
$$

where $G^{j+1}(x)$ is the cdf of RL distribution with a power parameter $j+1$. In other words, the cdf of EGRL distribution can be defined as a linear combination of the cdfs of RL distributions. By taking the derivative of $G^{j+1}(x)$ with respect to $x \geq 0$, we obtain the linear representation of the pdf of EGRL distribution which is given by

$$
f(x)=\sum_{j=0}^{\infty} \sum_{k=1}^{\infty}(-1)^{j+k+1}\binom{\beta}{k}\binom{\alpha k}{j+1}(j+1) g(x) G^{j}(x)
$$

where $g(x)$ is the pdf and $G(x)$ is the cdf of the baseline RL distribution.

## 3. Parameter Estimation

In this section, we present the parameter estimation procedure by means of four methods: maximum likelihood estimation (MLE), least squares estimation (LSE), weighted least squares estimation (WLSE), and Cramer von Mises estimation (CVME). The LSE, WLSE, and CVME methods are included in the study as an alternative to MLE due to their ease of use.
3.1. Maximum likelihood estimation. The $\log$ likelihood function of a random sample $X=\left(X_{1}, X_{2}, \ldots, X_{n}\right)$ from the EGRL distribution is given by

$$
\begin{align*}
\ell & =n \log (\alpha)+n \log (\beta)+n \log (\theta)-n \alpha \log (1-\theta)+\sum_{i=1}^{n} \log \left(\theta^{2} x_{i}-2 \theta+1\right) \\
& -\alpha \theta \sum_{i=1}^{n} x_{i}+(\alpha-1) \sum_{i=1}^{n} \log \left(\theta^{2} x_{i}-\theta+1\right) \\
& +(\beta-1) \sum_{i=1}^{n} \log \left(1-\left[\frac{\left(\theta^{2} x_{i}-\theta+1\right)}{1-\theta} e^{-\theta x_{i}}\right]^{\alpha}\right) \tag{29}
\end{align*}
$$

The elements of score vector $\left(\frac{\partial \ell}{\partial \alpha}, \frac{\partial \ell}{\partial \beta}, \frac{\partial \ell}{\partial \theta}\right)$ containing the first derivatives (stated otherwise the gradients) of the log likelihood function with respect to parameters $\alpha, \beta$ and $\theta$ are given below.

$$
\begin{align*}
\frac{\partial \ell}{\partial \alpha} & =\frac{n}{\alpha}-n \log (1-\theta)-\theta \sum_{i=1}^{n} x_{i}+\sum_{i=1}^{n} \log \left(\theta^{2} x_{i}-\theta+1\right) \\
& +(1-\beta) \sum_{i=1}^{n} \frac{\zeta_{i}^{\alpha}\left(\log \left(\zeta_{i}\right)-\theta x_{i}\right)}{e^{\alpha \theta x_{i}}-\zeta_{i}^{\alpha}} \\
\frac{\partial \ell}{\partial \beta} & =\frac{n}{\beta}+\sum_{i=1}^{n} \log \left(1-\left(\zeta_{i} e^{-\theta x_{i}}\right)^{\alpha}\right) \\
\frac{\partial \ell}{\partial \theta} & =\frac{n}{\theta}+\frac{n \alpha}{1-\theta}+\sum_{i=1}^{n} \frac{2\left(\theta x_{i}-1\right)}{\theta^{2} x_{i}-2 \theta+1}-\alpha \sum_{i=1}^{n} x_{i}+(\alpha-1) \sum_{i=1}^{n} \frac{2 \theta x_{i}-1}{\theta^{2} x_{i}-\theta+1} \\
& +(1-\beta) \sum_{i=1}^{n} \frac{\alpha x_{i} \zeta_{i}^{\alpha}\left[\theta^{3} x_{i}-\theta^{2}\left(x_{i}+2\right)+4 \theta-1\right]}{(1-\theta)\left(\theta^{2} x_{i}-\theta+1\right)\left(e^{\alpha \theta x_{i}}-\zeta_{i}^{\alpha}\right)} \tag{30}
\end{align*}
$$

where $\zeta_{i}=\frac{\left(\theta^{2} x_{i}-\theta+1\right)}{1-\theta}$.
Maximum likelihood estimates (MLEs) are described analytically by setting the elements of the score vector equal to zero and solving for each parameter. The resulting equations $\frac{\partial \ell}{\partial \alpha}=0, \frac{\partial \ell}{\partial \beta}=0$ and $\frac{\partial \ell}{\partial \theta}=0$ need to be solved simultaneously. Maximizing the log likelihood function with respect to parameters $\alpha, \beta$ and $\theta$ can be performed using a reliable non-linear optimization technique such as Nelder and

Mead (NM) or the Broyden-Fletcher-Goldfarb-Shanno (BFGS) algorithm. Maximizing the log likelihood (or minimizing minus the log likelihood) function can be achieved by using maxLik and optim subroutines of maxLik and stats packages in R statistical software.

When maximizing the $\log$ likelihood function above, the initials of $\alpha, \beta$, and $\theta$ parameters must be specified. To easily obtain the initial for parameter $\theta$ by means of using the usual RL distribution in equation (1), the initials of parameters $\alpha$ and $\beta$ were set to 1 . The initial of parameter $\theta$ was obtained by taking the inverse of the root obtained in $\mu=\frac{\lambda^{2}}{\lambda-1}$ (see 24 p. 250]), where parameter $\mu$ was replaced by the sample mean $\bar{x}$. The resulting initial for this parameter is $\theta_{\text {init }}=\frac{2}{\bar{x}+\sqrt{\bar{x}^{2}-4 \bar{x}}}$.

We do not provide the analytical expressions of the entries of Hessian matrix for the log likelihood function of EGRL distribution which are too complicated. The standard errors of model parameters can be obtained by an approximate Hessian matrix using the default option (i.e., the finite-difference approach) in the maxLik package. The square root of diagonals for the inverse of minus the Hessian matrix gives the standard errors of parameter estimates. However, the same standard errors can be obtained by using summary function in the maxLik package. Note that this approximation technique does not always converge for the standard errors of model parameters. In such a case, nonparametric bootstrapping (NB; 10 ) is a reasonable alternative to estimate the standard errors of parameters. The NB can also be used to obtain an estimate of bias to compare the performance of estimation methods presented in this paper, which will be evaluated in the application section.

The estimates of model parameters and their standard errors can also be obtained by maximizing a function of the cdf of EGRL distribution or a weighted form of this function known as the method of (weighted) least squares estimation which will be presented in the next subsection.
3.2. The method of (Weighted) Least-squares estimation. Based on Swain et al. ( 33 ), the least squares estimates of model parameters and their standard errors can be attained by maximizing

$$
\begin{equation*}
-\sum_{i=1}^{n}\left[F\left(x_{(i)}\right)-\frac{i}{n+1}\right]^{2}, \tag{31}
\end{equation*}
$$

where $F\left(x_{(i)}\right)$ is the cdf of the ordered random variables $x_{(1)}<x_{(2)}<\ldots<x_{(n)}$, see also [28, p. 181]. Thus, the least squares estimates for the EGRL distribution are obtained by maximizing

$$
\begin{equation*}
-\sum_{i=1}^{n}\left[\left(1-\left[1-G\left(x_{(i)}\right)\right]^{\alpha}\right)^{\beta}-\frac{i}{n+1}\right]^{2} \tag{32}
\end{equation*}
$$

where $F\left(x_{(i)}\right)$ in equation (33) is replaced by the cdf of the ordered random variables for the EGRL distribution. Here, $G\left(x_{(i)}\right)$ represents the cdf of the ordered random variables for the baseline RL distribution in equation (2).

The weighted least squares estimation (WLSE) can be more reliable than the usual least squares estimation (LSE) when the data involve heteroscedasticity which often occurs in the presence of outlier(s). The WLSE incorporates an additional weight factor into the function above to quantify the importance of each observation in the data when estimating model parameters. The WLSE is (often) less sensitive to outliers when compared to the usual LSE ${ }^{1}$. The weighted least squares estimates can be obtained by maximizing

$$
\begin{equation*}
-\sum_{i=1}^{n} w_{(i)}\left[F\left(x_{(i)}\right)-\frac{i}{n+1}\right]^{2} \tag{33}
\end{equation*}
$$

where $w_{(i)}=\frac{(n+1)^{2}(n+2)}{i(n-i+1)}$ is the value of weight factor of the $i$ th observation for the data in increasing order, 28, p. 181]. Similar to the least squares estimates, the weighted least squares estimates for the EGRL distribution are obtained by maximizing

$$
\begin{equation*}
-\sum_{i=1}^{n} w_{(i)}\left[\left(1-\left[1-G\left(x_{(i)}\right)\right]^{\alpha}\right)^{\beta}-\frac{i}{n+1}\right]^{2} \tag{34}
\end{equation*}
$$

Another popular estimation method that can easily be applied to estimate model parameters for the EGRL distribution is the method of Cramer-Von-Mises estimation (CVME), which will be detailed in the next subsection.
3.3. The method of Cramer-von-Mises estimation. The estimates of model parameters using the Cramer-von-Mises estimation (CVME; 20) is obtained by maximizing another function of the cdf of EGRL distribution which is given by

$$
\begin{equation*}
-\frac{1}{12 n}-\sum_{i=1}^{n}\left[F\left(x_{(i)}\right)-\frac{2 i-1}{2 n}\right]^{2} . \tag{35}
\end{equation*}
$$

Similar to equations (34) and (36), this function for the EGRL distribution can be defined as

$$
\begin{equation*}
-\frac{1}{12 n}-\sum_{i=1}^{n}\left[\left(1-\left[1-G\left(x_{(i)}\right)\right]^{\alpha}\right)^{\beta}-\frac{2 i-1}{2 n}\right]^{2} \tag{36}
\end{equation*}
$$

The gradients and analytical expressions of Hessian matrices with respect to the maximized functions using LSE, WLSE, and CVME methods are not presented here, but, would be made available upon request.

## 4. Simulation Studies

This section presents two simulation studies, first of which aims to investigate the performances of MLE, LSE, WLSE, and CVME methods for the EGRL distribution with respect to the bias, precision, and accuracy measures given in Walther and Moore ( 34 ). The second simulation study is however set up to illustrate

[^6]the potentiality of the new EGRL distribution in comparison to the some other lifetime distributions listed in Table 2. In this simulation, we show that model

Table 2. Some selected lifetime distributions.

| Distribution | Author(s) |
| :--- | :--- |
| Rayleigh | Rayleigh ( 25) |
| Exponentiated generalized Normal (EGN) | Corderio et al. ( 7 ) |
| Exponentiated generalized Gumbel (EGGu) | Corderio et al. ( 7 ) |
| Exponentiated generalized Ramos-Louzada (EGRL) | (New) |

fit indices should be used in conjunction with information criteria to detect the best distribution in a set of distributions when analyzing the data. For the selection of best fitting models, the Cramer-von Mises (C*; 9 ), Watson ( $\mathrm{W}^{*}$; 35 ), Kuiper ( $\mathrm{K}^{*}$; 17 ) and Kolmogorov-Smirnov (KS*; 16 31) goodness of fit statistics and the $\log$ likelihood ( $\ell$ ), Akaike information criterion (AIC; 1, 2), Consistent Akaike information criterion (CAIC; 6), Corrected Akaike information criterion (AICc; 14), Bayesian information criterion (BIC; 29 ), and Hannan-Quinn information criterion (HQIC; 12) are used. The goodness of fit statistics are given by

$$
\begin{align*}
C^{*} & =\frac{1}{12 n}+\sum_{k=1}^{n}\left[\frac{2 k-1}{2 n}-F\left(X_{(k)}\right)\right]^{2}, \\
\mathrm{~W}^{*} & =\sqrt{C^{*}-n\left(\left[\frac{1}{n} \sum_{k=1}^{n} F\left(X_{(k)}\right)\right]-\frac{1}{2}\right)^{2}}, \\
\mathrm{~K}^{*} & =\max \left[\frac{k}{n}-F\left(x_{(k)}\right)\right]+\max \left[F\left(X_{(k)}\right)-\frac{k-1}{n}\right], \\
\mathrm{KS}^{*} & =\max \left[F\left(X_{(k)}\right)-\frac{k-1}{n}, \frac{k}{n}-F\left(X_{(k)}\right)\right], \tag{37}
\end{align*}
$$

where $n$ is the sample size and $F(x)$ is the cdf of the distribution under consideration for which the values of random variable X are in increasing order, namely, $x_{(1)}<$ $x_{(2)}<\ldots<x_{(n)}$. The small values of these information criteria and test statistics above imply better model fits. Similarly, the information criteria are given by

$$
\begin{aligned}
A I C & =-2 \ell+2 p \\
\mathrm{AICc} & =-2 \ell+\frac{2 p n}{n-p-1} \\
\mathrm{CAIC} & =-2 \ell+p[\log (n)+1]
\end{aligned}
$$

$$
\begin{align*}
\mathrm{BIC} & =-2 \ell+p \log (n) \\
\mathrm{HQIC} & =-2 \ell+2 p \log [\log (n)] \tag{38}
\end{align*}
$$

where $n$ is the sample size and $p$ is the number of parameters in the model.
The simulation studies for two sets of population values of parameters for the EGRL distribution comprise the following steps.
(1) (a) For the first simulation: Set $\alpha=1, \beta=1$, and $\theta=0.3$ as the population values of parameters for the EGRL distribution.
(b) For the second simulation: Set $\alpha=1.2, \beta=1.3$, and $\theta=0.3$ as the population values of parameters for the EGRL distribution.
(2) Set the sample size as $N=20,100$, or 500 .
(3) Generate the values of EGRL distribution based on the population values in Step 1 and the sample size in Step 2. The data generation from the EGRL distribution is performed by using the automatic nonuniform random variate generation process presented in Hörmann et al. ( 13 ). This procedure can easily be implemented using tdr. new and ur subroutines of Runuran package in $R$ statistical software.
(4) Obtain the estimates of model parameters using MLE, LSE, WLSE, and CVME methods for the EGRL distribution in the first simulation and for the distributions in Table 2 in the second simulation.
(5) Perform Steps 3-4 for $S=1000$ times.
(6) (a) For the first simulation: For the EGRL distribution and each estimation method, calculate the bias, precision, and accuracy measures given in Walther and Moore ( 34 ).
(b) For the second simulation: For each distribution and estimation method, calculate the log likelihood value, the values of goodness of fit statistics, and information criteria in equations (31), (39) and (40), respectively.
Notably, the measures in Step 6 (a) are obtained for each parameter of EGRL distribution in $S=1000$ simulations. For example, the bias, precision, and accuracy measures for parameter $\alpha$ are given by

$$
\begin{array}{r}
\operatorname{Bias}(\alpha)=\frac{1}{S} \sum_{s=1}^{S}\left(\hat{\alpha}_{s}-\alpha\right) \\
\operatorname{Precision}(\alpha)=\frac{1}{S} \sum_{s=1}^{S}\left(\hat{\alpha}_{s}-\bar{\alpha}\right)^{2} \\
\operatorname{Accuracy}(\alpha)=\frac{1}{S} \sum_{s=1}^{S}\left(\hat{\alpha}_{s}-\alpha\right)^{2} \tag{41}
\end{array}
$$

where $\alpha=1$ is the population value of $\alpha$ in Step 1 (a), $\hat{\alpha}_{s}$ is the estimate of parameter $\alpha$ in the $s$ th simulation, and $\bar{\alpha}=\frac{1}{S} \sum_{s=1}^{S} \hat{\alpha}_{s}$ for $s=1,2, \ldots, 1000$. Analogous calculations are performed for parameters $\beta$ and $\theta$. These values are displayed in

Table 3. Similarly, the values obtained in Step 6 (b) are presented in Tables 4, 5, 6 , and 7 in each of which one of the estimation methods concerned are displayed in turn.

In the first and second simulation studies, parameter estimation over the data sets was performed by NM and BFGS algorithms, respectively. A data set was not accepted for inclusion in $S=1000$ simulation trials if at least one of the following conditions occured. (1) The initial of parameter $\theta$ was not in the range of 0 and 0.5 or sample mean is smaller than 4 in line with $\theta_{\text {init }}=\frac{2}{\bar{x}+\sqrt{\bar{x}^{2}-4 \bar{x}}}$ (see page 11). (2) The estimates of parameters were obtained outside the parameter space. For example, when the estimate of parameter $\alpha>0$ for the EGN distribution is obtained as $\hat{\alpha}<0$. (3) When the convergence criterion was not obtained for any of the distributions under consideration. (4) When the log likelihood value for any distribution in the set was obtained as minus infinity. Note that this last condition only applies to the second simulation study. If at least one of the conditions above occurs in the simulation, a different data was generated for the corresponding simulation trial.

Table 3 shows the values of bias, precision, and accuracy measures for each parameter of the EGRL distribution obtained from $S=1000$ random datasets using MLE, LSE, WLSE, and CVME methods with NM algorithm. A small value in the table represents a small bias, a high precision, or a high accuracy measure. This table displays that increasing the sample size eventually reduces the bias and increases the precision and accuracy for parameters $\alpha, \beta$, and $\theta$ of the EGRL distribution using each estimation method. The performance of each estimation method improves as the sample size increases. It is concluded that MLE outperforms other estimation methods in terms of bias, precision, and accuracy measures.

Tables $4,5,6$, and 7 show the average values of the (minus) log likelihood, goodness of fit statistics, and information criteria for each distribution under evaluation using MLE, LSE, WLSE, and CVME methods with BFGS algorithm, respectively. Based on these tables, one-parameter Rayleigh distribution does not provide enough flexibility in modeling the data. Because model fit statistics and (minus) log likelihood values for this distribution are larger than other distributions. It seems that the EGRL distribution often has smaller, and thus, better model fit statistics and (minus) log likelihood values when compared to other distributions. However, note that, these goodness of fit statistics are biased themselves, since they do not take the model complexity into account when choosing the best distribution in a set of distributions. The information criteria like the AIC and BIC reduce this bias by penalizing model complexity (i.e., penalizing the models containing unnecessarily more parameters). For example, when estimating model parameters using MLE for $n=20$ in Table 4, the best distribution in the set according to the values of model fit statistics is the EGGu distribution, while it is the second best distribution after the EGRL distribution based on the values of all the information criteria under consideration. Sample size plays a crucial role for information criteria when detecting the best distribution in a set. Because small samples tend to support

TABLE 3. Bias, precision and accuracy measures for the parameters of EGRL distribution

more parsimonious (stated otherwise simple) models, while large samples tend to support more complex models. The performance of EGRL distribution increase better than that of other distributions in the set as the sample size increases. The EGRL distribution in these tables is associated with the smallest (average) values of information criteria for $n=500$, regardless of the type of information criterion or estimation method. Moreover, the EGRL distribution performs better than other exponentiated generalized distributions, namely, the EGN and EGGu distributions, in all cases where the sample size is $n=100$ or $n=500$.

Table 4. The values of the goodness of fit statistics, (minus) log likelihood, and information criteria for each distribution under evaluation using MLE method.

| Sample <br> size $(n)$ | Models | $\overline{\mathrm{C}^{*}}$ | $\overline{\mathrm{~W}^{*}}$ | $\overline{\mathrm{~K}^{*}}$ | $\overline{\mathrm{KS}^{*}}$ | $-\bar{\ell}$ | $\overline{\text { AIC }}$ | $\overline{\text { AICc }}$ | $\overline{\mathrm{CAIC}}$ | $\overline{\mathrm{BIC}}$ | $\overline{\text { HQIC }}$ |
| :--- | :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 20 | Rayleigh | 0.44 | 0.43 | 0.37 | 0.28 | 57.00 | 115.99 | 116.21 | 117.99 | 116.99 | 116.18 |
|  | EGRL | 0.06 | 0.23 | 0.23 | 0.14 | 51.33 | 108.67 | 111.04 | 114.66 | 111.66 | 109.25 |
|  | EGN | 0.12 | 0.30 | 0.28 | 0.17 | 54.69 | 117.38 | 119.12 | 125.36 | 121.36 | 118.16 |
|  | EGGu | 0.06 | 0.22 | 0.23 | 0.13 | 51.57 | 111.14 | 113.81 | 119.13 | 115.13 | 111.92 |
| 100 | Rayleigh | 1.69 | 0.79 | 0.27 | 0.22 | 276.01 | 554.02 | 554.07 | 557.63 | 556.63 | 555.08 |
|  | EGRL | 0.06 | 0.22 | 0.10 | 0.06 | 251.62 | 509.23 | 509.63 | 520.05 | 517.05 | 512.40 |
|  | EGN | 0.38 | 0.52 | 0.20 | 0.12 | 268.78 | 545.57 | 545.84 | 559.99 | 555.99 | 549.78 |
|  | EGGu | 0.07 | 0.25 | 0.11 | 0.07 | 253.58 | 515.16 | 515.59 | 529.58 | 525.58 | 519.38 |
| 500 | Rayleigh | 7.74 | 1.66 | 0.24 | 0.20 | 1369.52 | 2741.03 | 2741.04 | 2746.25 | 2745.25 | 2742.68 |
|  | EGRL | 0.05 | 0.21 | 0.05 | 0.03 | 1255.06 | 2516.11 | 2516.19 | 2531.76 | 2528.76 | 2521.08 |
|  | EGN | 1.33 | 0.97 | 0.16 | 0.09 | 1328.86 | 2665.72 | 2665.77 | 2686.58 | 2682.58 | 2672.33 |
|  | EGGu | 0.15 | 0.35 | 0.07 | 0.04 | 1265.08 | 2538.17 | 2538.25 | 2559.03 | 2555.03 | 2544.78 |

Table 5. The values of the goodness of fit statistics, (minus) log likelihood, and information criteria for each distribution under evaluation using LSE method.

| Sample <br> size $(n)$ | Models | $\overline{\mathrm{C}^{*}}$ | $\overline{\mathrm{~W}^{*}}$ | $\overline{\mathrm{~K}^{*}}$ | $\overline{\mathrm{KS}^{*}}$ | $\bar{\ell}$ | $\overline{\mathrm{AIC}}$ | $\overline{\mathrm{AICc}}$ | $\overline{\mathrm{CAIC}}$ | $\overline{\mathrm{BIC}}$ | $\overline{\mathrm{HQIC}}$ |
| :--- | :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 20 | Rayleigh | 0.17 | 0.38 | 0.34 | 0.20 | 59.43 | 120.85 | 121.08 | 122.85 | 121.85 | 121.05 |
|  | EGRL | 0.04 | 0.21 | 0.22 | 0.12 | 51.14 | 108.27 | 110.65 | 114.26 | 111.26 | 108.86 |
|  | EGN | 0.06 | 0.23 | 0.24 | 0.13 | 60.90 | 129.81 | 131.54 | 137.79 | 133.79 | 130.59 |
|  | EGGu | 0.04 | 0.19 | 0.20 | 0.11 | 52.29 | 112.58 | 115.25 | 120.57 | 116.57 | 113.36 |
| 100 | Rayleigh | 0.63 | 0.77 | 0.26 | 0.14 | 288.01 | 578.01 | 578.05 | 581.62 | 580.62 | 579.07 |
|  | EGRL | 0.04 | 0.19 | 0.10 | 0.05 | 251.20 | 508.39 | 508.79 | 519.21 | 516.21 | 511.56 |
|  | EGN | 0.14 | 0.35 | 0.16 | 0.10 | 285.02 | 578.05 | 578.32 | 592.47 | 588.47 | 582.26 |
|  | EGGu | 0.04 | 0.18 | 0.09 | 0.05 | 256.61 | 521.23 | 521.65 | 535.65 | 531.65 | 525.44 |
| 500 | Rayleigh | 2.89 | 1.68 | 0.23 | 0.12 | 1437.51 | 2877.03 | 2877.03 | 2882.24 | 2881.24 | 2878.68 |
|  | EGRL | 0.04 | 0.18 | 0.04 | 0.02 | 1255.46 | 2516.92 | 2516.99 | 2532.56 | 2529.56 | 2521.88 |
|  | EGN | 0.54 | 0.70 | 0.13 | 0.08 | 1391.44 | 2790.87 | 2790.92 | 2811.73 | 2807.73 | 2797.49 |
|  | EGGu | 0.04 | 0.20 | 0.05 | 0.03 | 1286.80 | 2581.61 | 2581.69 | 2602.46 | 2598.46 | 2588.22 |

## 5. Application

This data contain $N=116$ observations representing a mean ozone in parts per billion at Rosevelt Island. These observations are obtained from the airquality dataset in datasets package of $R$ statistical software (version 4.2.2). Table 8 shows

Table 6. The values of the goodness of fit statistics, (minus) log likelihood, and information criteria for each distribution under evaluation using WLSE method.

| Sample <br> size $(n)$ | Models | $\overline{\mathrm{C}^{*}}$ | $\overline{\mathrm{~W}^{*}}$ | $\overline{\mathrm{~K}^{*}}$ | $\overline{\mathrm{KS}^{*}}$ | $-\bar{\ell}$ | $\overline{\mathrm{AIC}}$ | $\overline{\mathrm{AICc}}$ | $\overline{\mathrm{CAIC}}$ | $\overline{\mathrm{BIC}}$ | $\overline{\mathrm{HQIC}}$ |
| :--- | :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 20 | Rayleigh | 0.17 | 0.38 | 0.33 | 0.20 | 59.04 | 120.09 | 120.31 | 122.08 | 121.08 | 120.28 |
|  | EGRL | 0.05 | 0.21 | 0.22 | 0.12 | 50.92 | 107.83 | 110.21 | 113.82 | 110.82 | 108.42 |
|  | EGN | 0.06 | 0.24 | 0.24 | 0.13 | 60.61 | 129.22 | 130.96 | 137.21 | 133.21 | 130.00 |
|  | EGGu | 0.04 | 0.19 | 0.20 | 0.11 | 52.24 | 112.48 | 115.15 | 120.47 | 116.47 | 113.26 |
| 100 | Rayleigh | 0.63 | 0.76 | 0.26 | 0.15 | 286.51 | 575.02 | 575.06 | 578.62 | 577.62 | 576.07 |
|  | EGRL | 0.04 | 0.20 | 0.10 | 0.05 | 250.92 | 507.83 | 508.23 | 518.65 | 515.65 | 510.99 |
|  | EGN | 0.16 | 0.38 | 0.16 | 0.09 | 372.13 | 752.26 | 752.53 | 766.68 | 762.68 | 756.48 |
|  | EGGu | 0.04 | 0.18 | 0.09 | 0.05 | 256.45 | 520.91 | 521.33 | 535.33 | 531.33 | 525.12 |
| 500 | Rayleigh | 2.91 | 1.66 | 0.23 | 0.12 | 1428.25 | 2858.50 | 2858.51 | 2863.71 | 2862.71 | 2860.15 |
|  | EGRL | 0.04 | 0.19 | 0.04 | 0.02 | 1255.35 | 2516.69 | 2516.77 | 2532.34 | 2529.34 | 2521.65 |
|  | EGN | 0.56 | 0.72 | 0.15 | 0.10 | 3498.90 | 7005.81 | 7005.86 | 7026.67 | 7022.67 | 7012.42 |
|  | EGGu | 0.04 | 0.20 | 0.05 | 0.03 | 1286.84 | 2581.68 | 2581.76 | 2602.54 | 2598.54 | 2588.29 |

Table 7. The values of the goodness of fit statistics, (minus) log likelihood, and information criteria for each distribution under evaluation using CVME method.

| Sample <br> size $(n)$ | Models | $\overline{\mathrm{C}^{*}}$ | $\overline{\mathrm{~W}^{*}}$ | $\overline{\mathrm{~K}^{*}}$ | $\overline{\mathrm{KS}^{*}}$ | $\bar{\ell}$ | $\overline{\mathrm{AIC}}$ | $\overline{\mathrm{AICc}}$ | $\overline{\mathrm{CAIC}}$ | $\overline{\mathrm{BIC}}$ | $\overline{\mathrm{HQIC}}$ |
| :--- | :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 20 | Rayleigh | 0.18 | 0.39 | 0.34 | 0.20 | 59.83 | 121.66 | 121.88 | 123.65 | 122.65 | 121.85 |
|  | EGRL | 0.04 | 0.20 | 0.21 | 0.11 | 51.28 | 108.55 | 110.93 | 114.54 | 111.54 | 109.14 |
|  | EGN | 0.05 | 0.22 | 0.23 | 0.13 | 62.26 | 132.52 | 134.25 | 140.50 | 136.50 | 133.30 |
|  | EGGu | 0.03 | 0.18 | 0.19 | 0.10 | 52.73 | 113.45 | 116.12 | 121.44 | 117.44 | 114.23 |
| 100 | Rayleigh | 0.63 | 0.77 | 0.26 | 0.14 | 288.11 | 578.21 | 578.25 | 581.82 | 580.82 | 579.27 |
|  | EGRL | 0.04 | 0.19 | 0.09 | 0.05 | 251.28 | 508.57 | 508.96 | 519.38 | 516.38 | 511.73 |
|  | EGN | 0.14 | 0.35 | 0.16 | 0.09 | 283.67 | 575.33 | 575.61 | 589.75 | 585.75 | 579.55 |
|  | EGGu | 0.04 | 0.18 | 0.09 | 0.05 | 257.24 | 522.47 | 522.90 | 536.90 | 532.90 | 526.69 |
| 500 | Rayleigh | 2.89 | 1.68 | 0.23 | 0.12 | 1437.64 | 2877.29 | 2877.29 | 2882.50 | 2881.50 | 2878.94 |
|  | EGRL | 0.04 | 0.18 | 0.04 | 0.02 | 1255.44 | 2516.88 | 2516.95 | 2532.52 | 2529.52 | 2521.84 |
|  | EGN | 0.54 | 0.70 | 0.13 | 0.08 | 1391.87 | 2791.74 | 2791.79 | 2812.60 | 2808.60 | 2798.36 |
|  | EGGu | 0.04 | 0.20 | 0.05 | 0.02 | 1286.81 | 2581.61 | 2581.69 | 2602.47 | 2598.47 | 2588.23 |

the data and its descriptives. This dataset is heavily right skewed. The Q-Q plot in Figure 3 and Shapiro-Wilk normality test results ( $W=0.879, p<0.001$ ) show that the dataset is not normally distributed. The boxplot in Figure 3 displays that the dataset contains outliers. Table 9 shows the estimates of model parameters for the Ozone data using each of the estimation methods. We provide the R code on how
to obtain the estimates of model parameters using MLE in Appendix. The $R$ code for other estimation methods and distributions are not presented in Appendix, but, would be made available upon request.

Table 8. The Ozone data and its descriptives.


Table 9. The estimates of model parameters using MLE, LSE, WLSE, and CVME for the Ozone data.

|  | MLE |  |  |  | LSE |  |  |  |  |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| Models | $\alpha$ | $\beta$ | $\mu$ | $\theta$ | $\alpha$ | $\beta$ | $\mu$ | $\theta$ |  |
| Rayleigh | 37.774 | - | - | - | 28.295 | - | - | - |  |
| EGRL | 1.423 | 1.795 | - | 0.024 | 1.061 | 1.514 | - | 0.030 |  |
| EGN | 1.542 | 123.522 | -114.284 | 84.077 | 3.237 | 115.191 | -56.913 | 111.820 |  |
| EGGu | 0.182 | 0.602 | 16.227 | 7.177 | 0.115 | 0.648 | 12.322 | 5.188 |  |
|  |  | WLSE |  |  |  |  | CVM |  |  |
| Models | $\alpha$ | $\beta$ | $\mu$ | $\theta$ | $\alpha$ | $\beta$ | $\mu$ | $\theta$ |  |
| Rayleigh | 28.735 | - | - | - | 28.288 | - | - | - |  |
| EGRL | 2.999 | 1.654 | - | 0.011 | 1.893 | 1.551 | - | 0.017 |  |
| EGN | 0.452 | 218.532 | -179.939 | 47.622 | 3.306 | 109.547 | -51.712 | 110.068 |  |
| EGGu | 0.131 | 0.761 | 10.899 | 5.250 | 0.114 | 0.661 | 12.198 | 5.031 |  |

Figure 4 shows the pdfs, cdfs and survival and hazard rate functions for each distribution using MLE. This figure shows that the EGRL and EGGu distributions fit the data better than the Rayleigh and EGN distributions. Table 10 shows that the distribution of the observed Ozone data does not deviate significantly from the EGRL and EGGu distributions, but the distribution of the data differs from the Rayleigh and EGN distributions. This can be tested by the values of KolmogorovSmirnov (KS*) test statistics. For doing this, the critical value for the KS test is determined for $\alpha=0.05$, that is, $K S_{t}=\frac{1.36}{\sqrt{n}}=\frac{1.36}{\sqrt{116}}=0.126$. Therefore, for example, $K S^{*}=0.085<K S_{t}=0.126$ and $K S^{*}=0.065<K S_{t}=0.126$ indicate that the distribution of the Ozone data is not significantly different from the EGRL


Figure 3. The Q-Q plot and box plot for the Ozone data.


Figure 4. The pdf, cdf, survival and hazard rate functions for the Ozone data.
and EGGu distributions, respectively, when parameter estimation is performed by MLE. However, $K S^{*}=0.248>K S_{t}=0.126$ and $K S^{*}=0.214>K S_{t}=0.126$ mean that the distribution of the data significantly different from the Rayleigh and EGN distributions, respectively, when parameter estimation is performed by MLE. Table 10 also shows that the EGGu distribution provides the smallest goodness of fit statistics, regardless of the method used for parameter estimation. However, as noted in the introduction, these goodness of fit statistisc are biased as they do not
take the model complexity into account. Information criteria reduce this bias by considering both the fit and complexity of the model being evaluated. In Table 10, the values of information criteria indicate that the EGRL distribution has a better balance between the model fit and complexity when compared to Rayleigh, EGN, and EGGu distributions. Thus, we conclude that the EGRL distribution can be considered as an alternative distribution in the exponential generalized family of distributions when analyzing positively skewed data using MLE, LSE, WLSE, and CVME for parameter estimation.

Table 10. The values of the goodness of fit statistics, (minus) log likelihood, and information criteria for each distribution under evaluation using MLE, LSE, WLSE, and CVME for the Ozone data.

|  |  |  |  | MLE |  |  |  |  |  |  |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| Models | $\mathrm{C}^{*}$ | $\mathrm{~W}^{*}$ | $\mathrm{~K}^{*}$ | $\mathrm{KS}^{*}$ | $-\ell$ | AIC | AICc | CAIC | BIC | HQIC |
| Rayleigh | 2.17 | 1.03 | 0.31 | 0.25 | 561.99 | 1125.97 | 1126.01 | 1129.73 | 1128.73 | 1127.09 |
| EGRL | 0.12 | 0.33 | 0.14 | 0.09 | 541.40 | 1088.79 | 1089.01 | 1100.05 | 1097.05 | 1092.15 |
| EGN | 0.51 | 0.66 | 0.21 | 0.14 | 556.53 | 1121.06 | 1121.42 | 1136.08 | 1132.08 | 1125.53 |
| EGGu | 0.06 | 0.24 | 0.12 | 0.07 | 540.46 | 1088.91 | 1089.27 | 1103.93 | 1099.93 | 1093.39 |
|  |  |  |  |  | LSE |  |  |  |  |  |
| Models | $\mathrm{C}^{*}$ | $\mathrm{~W}^{*}$ | $\mathrm{~K}^{*}$ | $\mathrm{KS}^{*}$ | $-\ell$ | AIC | AICc | CAIC | BIC | HQIC |
| Rayleigh | 0.87 | 0.93 | 0.31 | 0.18 | 585.69 | 1173.38 | 1173.41 | 1177.13 | 1176.13 | 1174.50 |
| EGRL | 0.08 | 0.28 | 0.12 | 0.06 | 542.15 | 1090.30 | 1090.52 | 1101.56 | 1098.56 | 1093.65 |
| EGN | 0.32 | 0.55 | 0.20 | 0.11 | 558.28 | 1124.56 | 1124.92 | 1139.57 | 1135.57 | 1129.03 |
| EGGu | 0.03 | 0.18 | 0.08 | 0.04 | 540.99 | 1089.98 | 1090.34 | 1105.00 | 1101.00 | 1094.45 |
|  |  |  |  |  |  | WLSE |  |  |  |  |
| Models | $\mathrm{C}^{*}$ | $\mathrm{~W}^{*}$ | $\mathrm{~K}^{*}$ | $\mathrm{KS}^{*}$ | $-\ell$ | AIC | AICc | CAIC | BIC | HQIC |
| Rayleigh | 0.87 | 0.94 | 0.31 | 0.17 | 582.99 | 1167.98 | 1168.01 | 1171.73 | 1170.73 | 1169.10 |
| EGRL | 0.09 | 0.30 | 0.13 | 0.07 | 541.58 | 1089.15 | 1089.37 | 1100.41 | 1097.41 | 1092.51 |
| EGN | 0.39 | 0.62 | 0.22 | 0.12 | 556.00 | 1120.00 | 1120.35 | 1135.01 | 1131.01 | 1124.46 |
| EGGu | 0.03 | 0.18 | 0.10 | 0.05 | 540.28 | 1088.56 | 1088.92 | 1103.57 | 1099.57 | 1093.03 |
|  |  |  |  |  |  | CVME |  |  |  |  |
| Models | $\mathrm{C}^{*}$ | $\mathrm{~W}^{*}$ | $\mathrm{~K}^{*}$ | $\mathrm{KS}^{*}$ | $-\ell$ | AIC | AICc | CAIC | BIC | HQIC |
| Rayleigh | 0.87 | 0.93 | 0.31 | 0.18 | 585.74 | 1173.48 | 1173.51 | 1177.23 | 1176.23 | 1174.59 |
| EGRL | 0.08 | 0.28 | 0.12 | 0.06 | 541.96 | 1089.92 | 1090.14 | 1101.18 | 1098.18 | 1093.28 |
| EGN | 0.32 | 0.55 | 0.20 | 0.11 | 558.29 | 1124.59 | 1124.95 | 1139.60 | 1135.60 | 1129.06 |
| EGGu | 0.03 | 0.17 | 0.08 | 0.04 | 540.98 | 1089.96 | 1090.32 | 1104.98 | 1100.98 | 1094.43 |

Goodness of fit statistics and information criteria are originally created to compare the performance of models, but not to compare the performance of estimation methods. Therefore, we do not recommend using the results in Table 10 to compare
the performance of the methods in estimating the parameters of the EGRL distribution. For doing this, we used a bootstrap estimate of bias presented in Efron and Tibshirani (10).

Let $\eta=(\alpha, \beta, \theta)$ be the vector containing the parameters of EGRL distribution. Then, the bootstrap estimate of bias for each estimation method is calculated as follows:
(1) Create $B=1000$ bootstrap samples by resampling with replacement from the original data.
(2) Obtain the estimate of parameter vector $\eta$ for each of the bootstrap samples.
(3) Obtain the overall bootstrap estimate of parameter vector $\eta$, that is, $\eta^{*}$, by averaging the estimates among the bootstrap samples. The standard errors of parameter estimates (i.e., $\mathrm{SE}_{\mathrm{B}}^{*}$ ) are obtained by taking the square root of diagonals of the covariance matrix for the estimates in the bootstrap samples.
(4) Calculate the bootstrap estimate of bias for parameter vector $\eta$, that is, $\operatorname{Bias}_{\mathrm{B}}=\left|\eta^{*}-\hat{\eta}\right|$, where $\hat{\eta}$ is the vector of the usual estimates obtained for the original data using MLE, LSE, WLSE, or CVME.

Table 11 shows the performance evaluation of each estimation method in estimating the parameters of the EGRL distribution for the Ozone data. Efron and Tibshirani ( $\mid 10)$ state that the bias can be ignored if $\frac{\operatorname{Bias}_{B}}{S E_{B}^{*}} \leq 0.25$. Therefore, based on the results in Table 11, CVME outperforms other estimation methods for this particular example as it has the ratios smaller than 0.25 when estimating parameters $\alpha, \beta$, and $\theta$. In line with Efron and Tibshirani ( 10 ), the bias-adjusted estimates (BAEs) are also provided for each estimation method using $2 \hat{\eta}-\eta^{*}$. However, caution should be taken when using the bias-adjusted estimates in place of the usual estimates, as biases are more difficult to estimate than standard errors and correcting bias may produce higher standard errors [10, p. 138]. While the bias-adjusted estimates are reasonably close to the usual estimates for MLE and CVME, these estimates are not close to the usual estimates for LSE and WLSE. The WLSE even produces a negative bias-adjusted estimate for parameter $\theta$.

In summary, it is concluded that CVME performs better than other estimation methods in analyzing the Ozone data based on nonparametric bootstrapping bias assessment. In this sense, MLE also provides a reasonable set of parameter estimates. However, LSE and WLSE do not perform well when compared to MLE and CVME for analyzing the Ozone data using the EGRL distribution.

## 6. Discussion

In this study, we introduced a new distribution called the exponentiated generalized Ramos-Louzada distribution involving three parameters. We used four estimation methods (i.e., the MLE, LSE, WLSE, and CVME) for estimation. We assess the performance of these methods for the EGRL distribution by means of

TABLE 11. Performance evaluation of the estimation methods for the Ozone data in terms of the bias measure using nonparametric bootstrapping.

| Method | $\alpha$ | $\beta$ | $\theta$ |
| :---: | :---: | :---: | :---: |
| MLE ( $\hat{\eta}$ ) | 1.423 | 1.795 | 0.024 |
| NB ( $\eta^{*}$ ) | 1.016 | 1.830 | 0.046 |
| $\mathrm{SE}_{\mathrm{B}}^{*}$ | 0.411 | 0.231 | 0.035 |
| $\operatorname{Bias}_{\mathrm{B}}=\left\|\eta^{*}-\hat{\eta}\right\|$ | 0.407 | 0.035 | 0.022 |
| Ratio $=\frac{\text { Bias }^{\text {S }}}{\text { SE }}$ | 0.990 | 0.152 | 0.629 |
| $\mathrm{BAE}=2 \hat{\eta}-\eta^{*}$ | 1.830 | 1.760 | 0.002 |
| LSE ( $\hat{\eta}$ ) | 1.061 | 1.514 | 0.030 |
| NB ( $\eta^{*}$ ) | 1.783 | 1.564 | 0.020 |
| $\mathrm{SE}_{\mathrm{B}}^{*}$ | 0.431 | 0.225 | 0.012 |
| $\operatorname{Bias}_{\mathrm{B}}=\left\|\eta^{*}-\hat{\eta}\right\|$ | 0.722 | 0.050 | 0.010 |
| Ratio $=\frac{\text { Bias }^{*}}{\mathrm{SE}_{\mathrm{B}}^{*}}$ | 1.675 | 0.222 | 0.833 |
| $\mathrm{BAE}=2 \hat{\eta}-\eta^{*}$ | 0.339 | 1.464 | 0.040 |
| WLSE ( $\hat{\eta}$ ) | 2.999 | 1.654 | 0.011 |
| NB ( $\eta^{*}$ ) | 2.060 | 1.735 | 0.023 |
| $\mathrm{SE}_{\mathrm{B}}^{*}$ | 0.836 | 0.217 | 0.023 |
| $\operatorname{Bias}_{\mathrm{B}}=\left\|\eta^{*}-\hat{\eta}\right\|$ | 0.939 | 0.081 | 0.012 |
| Ratio $=\frac{\text { BiasB }^{\text {S }}}{\text { SE }}$ | 1.123 | 0.373 | 0.522 |
| $\mathrm{BAE}=2 \hat{\eta}-\eta^{*}$ | 3.938 | 1.573 | -0.001 |
| CVME ( $\hat{\eta}$ ) | 1.893 | 1.551 | 0.017 |
| NB ( $\eta^{*}$ ) | 1.792 | 1.602 | 0.020 |
| $\mathrm{SE}_{\mathrm{B}}^{*}$ | 0.439 | 0.233 | 0.013 |
| $\operatorname{Bias}_{\mathrm{B}}=\left\|\eta^{*}-\hat{\eta}\right\|$ | 0.101 | 0.051 | 0.003 |
| Ratio $=\frac{\text { Bias }^{\text {S }}}{\mathrm{SE}_{\mathrm{B}}}$ | 0.230 | 0.219 | 0.231 |
| $\mathrm{BAE}=2 \hat{\eta}-\eta^{*}$ | 1.994 | 1.500 | 0.014 |

using bias, precision, and accuracy measures, the goodness of fit statistics, and information criteria. To attain this objective, we first generate the datasets from the EGRL distribution in two simulations with varying values of sample size. Then, in the first simulation, we evaluate the performance of each estimation method for the EGRL distribution by means of using bias, precision, and accuracy measures. We obtain smaller bias and better precision and accuracy measures for each parameter of EGRL distribution as the sample size increases. It is concluded that MLE outperforms other estimation methods as the sample size increases. Second simulation study is conducted to evaluate the performance of a set of distributions for each estimation method separately. It is concluded that the performance of

EGRL distribution increase better than that of other distributions as the sample size increases when the data in fact follow the EGRL distribution.

The EGRL distribution is a flexible distribution that can be used to improve the model fit when compared to other exponentiated generalized distributions such as EGN and EGGu distributions. However, caution should be taken when using this distribution to analyze datasets in some certain circumstances. We are compelled to highlight two main limitations of the EGRL distribution when using it in conjunction with the estimation methods and information criteria presented in this paper. First, the performance of the EGRL distribution on modeling the data depends on the method utilized for estimation. For example, the methods presented in this paper might produce biased parameter estimates and their standard errors in the case of the data contain many missing values and/or outliers. In such cases, a different estimation method dealing with missing values and/or outliers better should be preferred over these estimation methods. Second, the AIC is prone to overfitting, that is, falsely choosing more complicated distributions containing more parameters over the simpler (stated otherwise more parsimonious) distributions for small samples. We do not suggest the use of the EGRL distribution for small samples, since it contains relatively more parameters when compared to the usual RL distribution. In the same sense, the AIC should not be used as a decision criterion when the set of distributions contains the EGRL distribution for small samples.

Author Contribution Statements Writing, programming, and analysis was performed by Yasin Altınışık. All authors read, commented, and approved the final article.

Declaration of Competing Interests The authors declare that they have no competing interest.

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## Appendix

The $R$ (version 4.2.2) code used to estimate the parameters of the EGRL distribution using MLE is given below.

```
library(maxLik)
set.seed(111)
# The Ozone data
xi <- c(41, 36, 12, 18, 28, 23, 19, 8, 7, 16, 11, 14, 18, 14, 34,
6, 30, 11, 1, 11, 4, 32, 23, 45, 115, 37, 29, 71, 39, 23, 21, 37,
20, 12, 13, 135, 49, 32, 64, 40, 77, 97, 97, 85, 10, 27, 7, 48, 35,
61, 79, 63, 16, 80, 108, 20, 52, 82, 50, 64, 59, 39, 9, 16, 78, 35,
66, 122, 89, 110, 44, 28, 65, 22, 59, 23, 31, 44, 21, 9, 45, 168, 73,
76, 118, 84, 85, 96, 78, 73, 91, 47, 32, 20, 23, 21, 24, 44, 21, 28,
9, 13, 46, 18, 13, 24, 16, 13, 23, 36, 7, 14, 30, 14, 18, 20)
# Sample size
n <- length(xi)
# Determining the initials for the parameters, respectively.
alphainit <- 1
betainit <- 1
thetainit<-2/(mean(xi)+sqrt(mean(xi)^2-4*mean(xi)))
# Maximizing the log likelihood function.
logLik <- function(param) {
alpha <- param[1]
beta <- param[2]
```

```
theta <- param[3]
# The four lines below are used to ensure that the estimates updated
# in the BFGS algorithm are in line with the parameter spaces.
if (alpha < 0) {alpha <- 0.0001}
if (beta < 0) {beta <- 0.0001}
if (theta < 0) {theta <- 0.0001}
if (theta > 0.5) {theta <- 0.5}
gx <- (((1+theta^2*xi-2*theta)*(theta))/(1-theta))*(exp(-theta*xi))
Gx <- 1-((1+theta^2*xi-theta)/(1-theta))*(exp(-theta*xi))
ll <- n*log(alpha)+n*log(beta)+sum(log(gx))+(alpha-1)*(sum(log(1-Gx)))+
(beta-1)*sum(log(1-((1-Gx)^(alpha))))
}
# Obtaining the results of the BFGS algorithm. Here,
# control = list(iterlim = 100000) is used to ensure successfull
# convergence of the BFGS algorithm.
model <- maxLik(logLik, start = c(alphainit, betainit, thetainit),
method = "BFGS", control = list(iterlim = 100000))
# Displaying the results
summary(model)
Maximum Likelihood estimation
BFGS maximization, 44 iterations
Return code 0: successful convergence
Log-Likelihood: -541.3966
3 free parameters
Estimates:
    Estimate Std. error t value Pr(> t)
[1,] 1.42313 2.40164 0.593 0.553
[2,] 1.79495 0.24685 7.271 3.56e-13 ***
[3,] 0.02424 0.04196 0.578
Signif. codes: 0 '***' 0.001 '**' 0.01 '*' 0.05 '.' 0.1 ' ' 1
```

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# PPF DEPENDENT COMMON FIXED POINTS OF GENERALIZED WEAKLY CONTRACTIVE TYPE MULTI-VALUED MAPPINGS 

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#### Abstract

In this paper, we introduce the notion of generalized weakly contractive type multi-valued mapping with respect to a single-valued mapping and prove the existence of PPF dependent coincidence points in Banach spaces. Further, we introduce the notion of generalized weakly contractive type multivalued mappings for a pair of multi-valued mappings and prove the existence of PPF dependent common fixed points in Banach spaces. We draw some corollaries and provide nontrivial examples to illustrate our results.


## 1. Introduction

The Banach contraction principle is one of the fundemental and useful result in fixed point theory and it plays an important role in solving problems related to non-linear functional analysis. In 1969, Nadler 20 extended Banach contraction principle to the context of set valued mapping. For more works on the existence of fixed points of multi-valued maps, we refer Kaneko 16 and Mizoguchi and Takahashi 19. In 1997, Alber and Gurre-Delabriere 1 introduced weakly contractive map which is a generalization of contraction map and obtained fixed point results in the setting of Hilbert spaces. Rhoades $\sqrt[22]{ }$ extended this concept to metric spaces and Bae 6 considered these type of multi-valued mappings. Bose and Roychowdhury 9.10 considered some generalized versions of these mappings and proved some fixed point theorems.

[^7]Let $(X, d)$ be a metric space and $K(X)$, the family of all non-empty compact subsets of $X$ and $H$ represents the Hausdorff distance induced by the metric d. i.e.,

$$
H(A, B)=\max \left\{\sup _{a \in A} d(a, B), \sup _{b \in B} d(A, b)\right\}
$$

for any $A, B \in K(X)$, where $d(a, B)=\inf _{b \in B} d(a, b)$ and $d(A, b)=\inf _{a \in A} d(a, b)$.
Definition 1. [6] A point $x \in X$ is said to be a fixed point of a multi-valued mapping $T: X \rightarrow K(X)$ if $x \in T x$.
Definition 2. A point $x \in X$ is said to be a coincidence point of two mappings $f, g: X \rightarrow X$ if $f(x)=g(x)$.
Definition 3. [9] A mapping $T: X \rightarrow X$ is said to be a generalized weakly contractive map with respect to $f: X \rightarrow X$ if

$$
\psi(d(T x, T y)) \leq \psi(d(f x, f y))-\phi(d(f x, f y))
$$

for all $x, y \in X$, where $\psi, \phi:[0, \infty) \rightarrow[0, \infty)$ are both continuous such that $\psi(t), \phi(t)>0$ for $t \in(0, \infty)$ and $\psi(0)=0=\phi(0)$. In addition, $\phi$ is non-decreasing and $\psi$ is monotonically increasing (strictly).

If $\psi(t)=t$ for all $t \in[0, \infty)$, and $f$ is the identity map in Definition 3 then we say that $T: X \rightarrow X$ is said to be a weakly contractive map.
Definition 4. [g] A multi-valued mapping $T: X \rightarrow K(X)$ is said to be a generalized weakly contractive map with respect to $f: X \rightarrow X$ if

$$
\psi(H(T x, T y)) \leq \psi(d(f x, f y))-\phi(d(f x, f y))
$$

for all $x, y \in X$, where $\psi, \phi:[0, \infty) \rightarrow[0, \infty)$ are both continuous such that $\psi(t), \phi(t)>0$ for $t \in(0, \infty)$ and $\psi(0)=0=\phi(0)$. In addition, $\phi$ is non-decreasing and $\psi$ is monotonically increasing (strictly).

If $f$ is the identity mapping then the multi-valued mapping $T: X \rightarrow K(X)$ is said to be generalized weakly contractive. If $\psi(t)=t$ for all $t \in[0, \infty)$, then the multi-valued mapping $T: X \rightarrow K(X)$ is said to be weakly contractive with respect to $f$.

In 1977, Bernfeld, Lakshmikantham and Reddy 8 introduced the concept of fixed point for mappings that have different domains and ranges which is called PPF (Past, Present and Future) dependent fixed point. Furthermore, they introduced the notation of Banach type contraction for a non-self mappings and proved the existence of PPF dependent fixed points of Banach type contractive mappings in the Razumikhin class. Several mathematicians proved the existence of PPF dependent fixed points of single-valued mapppings and multi-valued mappings, for more details we refer to $2-5,7,13,18$. In 2016, Farajzadeh, Kaewcharoen and Plubtieng 14 introduced the concept of PPF dependent fixed point of multi-valued mappings which is an extension of PPF dependent fixed point of single valued mapping and proved the existence of PPF dependent fixed point for multi-valued mappings.

Motivated by the research work of Bose and Roychowdhury 9 on weakly contractive maps, we extend the above said results for the case of PPF dependent coincidence points and PPF dependent common fixed points.

In this paper, we introduce the notion of generalized weakly contractive type multi-valued mapping with respect to a single-valued mapping and prove the existence of PPF dependent coincidence points in Banach spaces. Further, we introduce the notion of generalized weakly contractive type multi-valued mappings for a pair of multi-valued mappings and prove the existence of PPF dependent common fixed points in Banach spaces. We draw some corollaries and provide examples to illustrate our main results.

## 2. Preliminaries

In this paper, we denote the real line by $\mathbb{R}, \mathbb{R}^{+}=[0, \infty)$, the set of all natural numbers by $\mathbb{N}$. Let $\left(E,\|\cdot\|_{E}\right)$ be a Banach space and we denote it by simply by $E$. Let $I=[a, b] \subseteq \mathbb{R}$ and $E_{0}=C(I, E)$, the set of all continuous functions on $I$ equipped with the supremum norm $\|\cdot\|_{E_{0}}$ and we define it by $\|\phi\|_{E_{0}}=\sup _{a \leq t \leq b}\|\phi(t)\|_{E}$ for $\phi \in E_{0}$.

In our discussion, let $C B(E)$ be the collection of all non-empty closed and bounded subsets of $E$. Then the Hausdorff metric $H_{E}$ on $C B(E)$ induced by the norm $\|.\|_{E}$ is defined by

$$
H_{E}(A, B)=\max \left\{\sup _{a \in A} d(a, B), \sup _{b \in B} d(A, b)\right\}
$$

for any $A, B \in C B(E)$, where $d(a, B)=\inf _{b \in B}\|a-b\|_{E}$ and $d(A, b)=\inf _{a \in A}\|a-b\|_{E}$.
For a fixed $c \in I$, the Razumikhin class $R_{c}$ of functions in $E_{0}$ is defined by $R_{c}=\left\{\phi \in E_{0} \mid\|\phi\|_{E_{0}}=\|\phi(c)\|_{E}\right\}$ and $R_{c}(c)=\left\{\phi(c) \mid \phi \in R_{c}\right\}$. Clearly every constant function from $I$ to $E$ belongs to $R_{c}$ so that $R_{c}$ is a non-empty subset of $E_{0}$.
Definition 5. 8 Let $R_{c}$ be the Razumikhin class of continuous functions in $E_{0}$. Then, we say that
(i) the class $R_{c}$ is algebraically closed with respect to the difference if $\phi-\psi \in R_{c}$ whenever $\phi, \psi \in R_{c}$.
(ii) the class $R_{c}$ is topologically closed if it is closed with respect to the topology on $E_{0}$ by the norm $\|\cdot\|_{E_{0}}$.
The Razumikhin class of functions $R_{c}$ has the following properties.
Theorem 1. 2 Let $R_{c}$ be the Razumikhin class of functions in $E_{0}$. Then
(i) for any $\phi \in R_{c}$ and $\alpha \in \mathbb{R}$, we have $\alpha \phi \in R_{c}$.
(ii) the Razumikhin class $R_{c}$ is topologically closed with respect to the norm defined on $E_{0}$.
(iii) $\underset{c \in[a, b]}{\cap R_{c}}=\left\{\phi \in E_{0} \mid \phi: I \rightarrow E\right.$ is constant $\}$.

Definition 6. [8] Let $T: E_{0} \rightarrow E$ be a mapping. A function $\phi \in E_{0}$ is said to be a PPF dependent fixed point of $T$ if $T \phi=\phi(c)$ for some $c \in I$.

Definition 7. [8] Let $T: E_{0} \rightarrow E$ be a mapping. Then $T$ is called a Banach type contraction if there exists a constant $k \in[0,1)$ such that

$$
\|T \phi-T \psi\|_{E} \leq k\|\phi-\psi\|_{E_{0}}
$$

for any $\phi, \psi \in E_{0}$.
Theorem 2. 8 Let $T: E_{0} \rightarrow E$ be a Banach type contraction. Let $R_{c}$ be an algebraically closed with respect to the difference and topologically closed. Then, T has a unique PPF dependent fixed point in $R_{c}$.

Farajzadeh, Kaewcharoen and Plubtieng 14 introduced the concept of PPF dependent fixed points of multi-valued mappings as follows.
Definition 8. 14 Let $T: E_{0} \rightarrow C B(E)$ be a multi-valued mapping. A point $\phi \in E_{0}$ is said to be a PPF dependent fixed point of $T$ if $\phi(c) \in T \phi$ for some $c \in I$.

Definition 9. 14 Let $f: E_{0} \rightarrow E_{0}$ be a single-valued mapping and $T: E_{0} \rightarrow C B(E)$ be a multi-valued mapping. A point $\phi \in E_{0}$ is said to be a PPF dependent coincidence point of $f$ and $T$ if $f \phi(c) \in T \phi$ for some $c \in I$.

Here we observe that $f \phi$ is not a composition of $\phi$ and $f$.
Definition 10. [14 Let $S, T: E_{0} \rightarrow E$ be two single-valued mappings. A point $\phi \in E_{0}$ is said to be a PPF dependent common fixed point of $S$ and $T$ if $S \phi=T \phi=\phi(c)$ for some $c \in I$.

We denote

$$
\begin{gathered}
\Psi=\left\{\psi: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+} \mid \psi\right. \text { is continuous, monotonically increasing and } \\
\psi(t)=0 \Longleftrightarrow t=0\}
\end{gathered}
$$

and

$$
\Phi=\left\{\phi: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+} \mid \phi \text { is continuous and } \phi(t)=0 \Longleftrightarrow t=0\right\}
$$

We use the following results in our subsequent discussions.
Proposition 1. If $\left\{a_{n}\right\}$ and $\left\{b_{n}\right\}$ are two real sequences, $\left\{b_{n}\right\}$ is bounded, then $\liminf \left(a_{n}+b_{n}\right) \leq \liminf a_{n}+\limsup b_{n}$.

Lemma 1. 20 Let $A$ and $B$ be two non-empty compact subsets of a metric space $X$. If $a \in A$ then there exists $b \in B$ such that $d(a, b) \leq H(A, B)$.
Lemma 2. 3 Let $\left\{\phi_{n}\right\}$ be a sequence in $E_{0}$ such that $\left\|\phi_{n}-\phi_{n+1}\right\|_{E_{0}} \rightarrow 0$ as $n \rightarrow \infty$. If $\left\{\phi_{n}\right\}$ is not a Cauchy sequence, then there exists an $\epsilon>0$ and two subsequences $\left\{\phi_{m_{k}}\right\}$ and $\left\{\phi_{n_{k}}\right\}$ of $\left\{\phi_{n}\right\}$ with $m_{k}>n_{k}>k$ such that $\left\|\phi_{n_{k}}-\phi_{m_{k}}\right\|_{E_{0}} \geq \epsilon,\left\|\phi_{n_{k}}-\phi_{m_{k}-1}\right\|_{E_{0}}<\epsilon$ and
(i) $\lim _{k \rightarrow \infty}\left\|\phi_{n_{k}}-\phi_{m_{k}+1}\right\|_{E_{0}}=\epsilon$,
(ii) $\lim _{k \rightarrow \infty}\left\|\phi_{n_{k}+1}-\phi_{m_{k}}\right\|_{E_{0}}=\epsilon$,
(iii) $\lim _{k \rightarrow \infty}\left\|\phi_{n_{k}}-\phi_{m_{k}}\right\|_{E_{0}}=\epsilon$,
(iv) $\lim _{k \rightarrow \infty}\left\|\phi_{n_{k}+1}-\phi_{m_{k}+1}\right\|_{E_{0}}=\epsilon$.

## 3. Existence of PPF Dependent Coincidence Points

In this section, we introduce the concept of PPF dependent coincidence point of $f: E \rightarrow E$ and $T: E_{0} \rightarrow E$.

Definition 11. Let $f: E \rightarrow E$ and $T: E_{0} \rightarrow E$ be two mappings. A point $\phi \in E_{0}$ is said to be a PPF dependent coincidence point of $f$ and $T$ if $T \phi=(f \circ \phi)(c)$ for some $c \in I$, where $f \circ \phi$ denotes the composition of $\phi$ and $f$.

We observe that if $f$ is the identity mapping then PPF dependent coincidence point of $f$ and $T$ becomes PPF dependent fixed point of $T$.

Motivated by this idea, in the following, we now introduce the concept of PPF dependent coincidence point of $f: E \rightarrow E$ and $T: E_{0} \rightarrow C B(E)$.
Definition 12. Let $f: E \rightarrow E$ be a single-valued mapping and $T: E_{0} \rightarrow C B(E)$ be a multi-valued mapping. A point $\phi \in E_{0}$ is said to be a PPF dependent coincidence point of $f$ and $T$ if $(f \circ \phi)(c) \in T \phi$ for some $c \in I$, where $f \circ \phi$ denotes the composition of $\phi$ and $f$.

We observe that, if $f$ is an identity mapping then $\phi$ is a PPF dependent fixed point of the multi-valued mapping $T$.

Notation: Let $c \in I$. Let $f: E \rightarrow E$ and $\phi \in E_{0}$. We denote $(f \circ \phi)(c)$ by $f \phi(c)$.

In the following, we introduce the notion of generalized weakly contractive type multi-valued mappings.
Definition 13. Let $T: E_{0} \rightarrow C B(E)$. Let $f: E \rightarrow E$ be a continuous function. Then, $T$ is said to be a generalized weakly contractive type multi-valued mapping with respect to $f$ if there exist $\psi \in \Psi$ and $\phi \in \Phi$ such that

$$
\begin{equation*}
\psi\left(H_{E}(T \alpha, T \beta)\right) \leq \psi\left(\|f \alpha-f \beta\|_{E_{0}}\right)-\phi\left(\|f \alpha-f \beta\|_{E_{0}}\right) \tag{1}
\end{equation*}
$$

for any $\alpha, \beta \in E_{0}$.
We observe the following:
(i) if $f$ is the identity mapping in (1) then the mapping $T: E_{0} \rightarrow C B(E)$ is said to be generalized weakly contractive type multi-valued mapping;
(ii) if $\psi(t)=t$ for any $t \in \mathbb{R}^{+}$in (1) then the mapping $T: E_{0} \rightarrow C B(E)$ is said to be weakly contractive type multi-valued mapping with respect to $f$;
(iii) if both $f$ is the identity mapping and $\psi(t)=t$ for any $t \in \mathbb{R}^{+}$in (1) then the mapping $T: E_{0} \rightarrow C B(E)$ is said to be weakly contractive type multi-valued mapping.

Theorem 3. Let $T: E_{0} \rightarrow C B(E)$ and $f: E \rightarrow E$ be functions that satisfy the following conditions:
(i) $T$ is a generalized weakly contractive type multi-valued mapping with respect to $f$,
(ii) $T \phi \subseteq f\left(R_{c}\right)(c)=\left\{f \phi(c) \mid \phi \in R_{c}\right\}$ for any $\phi \in E_{0}$,
(iii) $R_{c}$ is algebraically closed with respect to the difference,
(iv) $f\left(R_{c}\right)$ is complete and
(v) $f\left(R_{c}\right) \subseteq R_{c}$.

Then, $T$ and $f$ have a PPF dependent coincidence point in $R_{c}$.
Proof. Let $\phi_{0} \in R_{c}$. Then, $T \phi_{0} \subseteq E$. Let $x_{1} \in E$ be such that $x_{1} \in T \phi_{0}$.
Since $T \phi_{0} \subseteq f\left(R_{c}\right)(c)$, we choose $\phi_{1}$ in $R_{c}$ such that $x_{1}=f \phi_{1}(c) \in T \phi_{0}$.
From (1), we have

$$
\psi\left(H_{E}\left(T \phi_{0}, T \phi_{1}\right)\right) \leq \psi\left(\left\|f \phi_{0}-f \phi_{1}\right\|_{E_{0}}\right)-\phi\left(\left\|f \phi_{0}-f \phi_{1}\right\|_{E_{0}}\right)
$$

Since $x_{1} \in T \phi_{0}$, by Lemma 1 there exists $x_{2} \in T \phi_{1}$ such that

$$
\begin{equation*}
\left\|x_{1}-x_{2}\right\|_{E} \leq H_{E}\left(T \phi_{0}, T \phi_{1}\right) \tag{2}
\end{equation*}
$$

Since $x_{2} \in T \phi_{1}$ and $T \phi_{1} \subseteq f\left(R_{c}\right)(c)$, we choose $\phi_{2}$ in $R_{c}$ such that $x_{2}=f \phi_{2}(c) \in T \phi_{1}$.
If $\phi_{1}=\phi_{2}$ then $\phi_{1}$ is a PPF dependent coincidence point of $f$ and $T$.
Suppose that $\phi_{1} \neq \phi_{2}$.
From (2), we have

$$
\left\|f \phi_{1}(c)-f \phi_{2}(c)\right\|_{E} \leq H_{E}\left(T \phi_{0}, T \phi_{1}\right)
$$

Since $R_{c}$ is algebraically closed with respect to the difference, we have

$$
\begin{equation*}
\left\|f \phi_{1}-f \phi_{2}\right\|_{E_{0}} \leq H_{E}\left(T \phi_{0}, T \phi_{1}\right) \tag{3}
\end{equation*}
$$

From (1), we have

$$
\psi\left(H_{E}\left(T \phi_{1}, T \phi_{2}\right)\right) \leq \psi\left(\left\|f \phi_{1}-f \phi_{2}\right\|_{E_{0}}\right)-\phi\left(\left\|f \phi_{1}-f \phi_{2}\right\|_{E_{0}}\right)
$$

Since $x_{2} \in T \phi_{1}$, by Lemma 1 there exists $x_{3} \in T \phi_{2}$ such that

$$
\begin{equation*}
\left\|x_{2}-x_{3}\right\|_{E} \leq H_{E}\left(T \phi_{1}, T \phi_{2}\right) \tag{4}
\end{equation*}
$$

Since $x_{3} \in T \phi_{2}$ and $T \phi_{2} \subseteq f\left(R_{c}\right)(c)$, we choose $\phi_{3}$ in $R_{c}$ such that $x_{3}=f \phi_{3}(c) \in T \phi_{2}$.
If $\phi_{2}=\phi_{3}$ then $\phi_{2}$ is a PPF dependent coincident point of $f$ and $T$.
Suppose that $\phi_{2} \neq \phi_{3}$.
From (4), we have

$$
\left\|f \phi_{2}(c)-f \phi_{3}(c)\right\|_{E} \leq H_{E}\left(T \phi_{1}, T \phi_{2}\right)
$$

Since $R_{c}$ is algebraically closed with respect to the difference, we have

$$
\begin{equation*}
\left\|f \phi_{2}-f \phi_{3}\right\|_{E_{0}} \leq H_{E}\left(T \phi_{1}, T \phi_{2}\right) \tag{5}
\end{equation*}
$$

On continuing this process, we get a sequence $\left\{f \phi_{n}\right\}$ in $R_{c}$ such that

$$
\begin{equation*}
x_{n}=f \phi_{n}(c) \in T \phi_{n-1},\left\|f \phi_{n}-f \phi_{n+1}\right\|_{E_{0}} \leq H_{E}\left(T \phi_{n-1}, T \phi_{n}\right) \text { for all } n \in \mathbb{N} \tag{6}
\end{equation*}
$$

Clearly,

$$
\begin{align*}
\psi\left(\| f \phi_{n}-\right. & \left.f \phi_{n+1} \|_{E_{0}}\right) \leq \psi\left(H_{E}\left(T \phi_{n-1}, T \phi_{n}\right)\right. \\
& \left.\leq \psi\left(\| f \phi_{n-1}-f \phi_{n}\right) \|_{E_{0}}\right)-\phi\left(\left\|f \phi_{n-1}-f \phi_{n}\right\|_{E_{0}}\right) \tag{7}
\end{align*}
$$

$$
<\psi\left(\left\|f \phi_{n-1}-f \phi_{n}\right\|_{E_{0}}\right) .
$$

Since $\psi$ is monotonically increasing function, we have

$$
\left\|f \phi_{n}-f \phi_{n+1}\right\|_{E_{0}} \leq\left\|f \phi_{n-1}-f \phi_{n}\right\|_{E_{0}} .
$$

Therefore, the sequence $\left\{\left\|f \phi_{n}-f \phi_{n+1}\right\|_{E_{0}}\right\}$ is a decreasing sequence in $\mathbb{R}^{+}$and hence it is convergent.
Let $\left\|f \phi_{n}-f \phi_{n+1}\right\|_{E_{0}} \rightarrow r$ as $n \rightarrow \infty$.
From (7), we have

$$
\psi\left(\left\|f \phi_{n}-f \phi_{n+1}\right\|_{E_{0}}\right) \leq \psi\left(\left\|f \phi_{n-1}-f \phi_{n}\right\|_{E_{0}}\right)-\phi\left(\left\|f \phi_{n-1}-f \phi_{n}\right\|_{E_{0}}\right) .
$$

On applying limits as $n \rightarrow \infty$ on both sides, we get
$\psi(r) \leq \psi(r)-\phi(r)$ and hence $r=0$.
Therefore,

$$
\begin{equation*}
\lim _{n \rightarrow \infty}\left\|f \phi_{n}-f \phi_{n+1}\right\|_{E_{0}}=0 . \tag{8}
\end{equation*}
$$

We now show that $\left\{f \phi_{n}\right\}$ is a Cauchy sequence.
Suppose that $\left\{f \phi_{n}\right\}$ is not a Cauchy sequence. Then, there exists an $\epsilon>0$ and two subsequences $\left\{f \phi_{m_{k}}\right\}$ and $\left\{f \phi_{n_{k}}\right\}$ of $\left\{f \phi_{n}\right\}$ such that for any $k \in \mathbb{N}, m_{k}>n_{k}>k$ such that

$$
\begin{equation*}
\left\|f \phi_{n_{k}}-f \phi_{m_{k}}\right\|_{E_{0}} \geq \epsilon . \tag{9}
\end{equation*}
$$

Let $m_{k}$ be the smallest positive integer greater than $n_{k}$ satisfying (9).
Then, $\left\|f \phi_{n_{k}}-f \phi_{m_{k}}\right\|_{E_{0}} \geq \epsilon$ and $\left\|f \phi_{n_{k}}-f \phi_{m_{k}-1}\right\|_{E_{0}}<\epsilon$.
By Lemma 2 we have
$\lim _{k \rightarrow \infty}\left\|f \phi_{n_{k}+1}-f \phi_{m_{k}}\right\|_{E_{0}}=\epsilon=\lim _{k \rightarrow \infty}\left\|f \phi_{n_{k}}-f \phi_{m_{k}+1}\right\|_{E_{0}}=\lim _{k \rightarrow \infty}\left\|f \phi_{n_{k}}-f \phi_{m_{k}}\right\|_{E_{0}}$. Now, we show that $\lim _{k \rightarrow \infty}\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}}=\epsilon$ for any $l_{1}, l_{2} \in \mathbb{N}$.
Let $l_{1}, l_{2} \in \mathbb{N}$. Now we consider

$$
\begin{aligned}
&\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}} \leq\left\|f \phi_{n_{k}+l_{1}}-f \phi_{n_{k_{k}}+l_{1}-1}\right\|_{E_{0}}+\left\|f \phi_{n_{k}+l_{1}-1}-f \phi_{n_{k}+l_{1-2}}\right\|_{E_{0}} \\
&+\ldots+\left\|f \phi_{n_{k}+1}-f \phi_{n_{k}}\right\|_{E_{0}}+\left\|f \phi_{n_{k}}-f \phi_{m_{k}+1}\right\|_{E_{0}} \\
&+\left\|f \phi_{m_{k}+1}-f \phi_{m_{k}+2}\right\|_{E_{0}}+\ldots+\left\|f \phi_{m_{k}+l_{2}-1}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}} .
\end{aligned}
$$

On applying limit superior as $k \rightarrow \infty$ on both sides, we get

$$
\begin{equation*}
\limsup _{k \rightarrow \infty}\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}} \leq \epsilon . \tag{10}
\end{equation*}
$$

Now, we consider

Now, by applying Proposition 1 with $a_{k}=\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}}$ and
$b_{k}=\left(| | f \phi_{n_{k}}-f \phi_{n_{k}+1}\left\|_{E_{0}}+\right\| f \phi_{n_{k}+1}-f \phi_{n_{k}+2}\left\|_{E_{0}}+\ldots+\right\| f \phi_{n_{k}+l_{1}-1}-f \phi_{n_{k}+l_{1}} \|_{E_{0}}+\right.$ $\left\|f \phi_{m_{k}+l_{2}}-f \phi_{m_{k}+l_{2}-1}\right\|_{E_{0}}+\ldots+\left\|f \phi_{m_{k}+2}-f \phi_{m_{k}+1}\right\|_{E_{0}}$ ) we have
$\epsilon \leq \liminf _{k \rightarrow \infty}\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}}+\limsup _{k \rightarrow \infty}\left(\left\|f \phi_{n_{k}}-f \phi_{n_{k}+1}\right\|_{E_{0}}\right.$
$+\left\|f \phi_{n_{k}+1}-f \phi_{n_{k}+2}\right\|_{E_{0}}+\ldots+\left\|f \phi_{n_{k}+l_{1}-1}-f \phi_{n_{k}+l_{1}}\right\|_{E_{0}}+\left\|f \phi_{m_{k}+l_{2}}-f \phi_{m_{k}+l_{2}-1}\right\|_{E_{0}}$ $\left.+\ldots+\left\|f \phi_{m_{k}+2}-f \phi_{m_{k}+1}\right\|_{E_{0}}\right)$.

Hence

$$
\begin{equation*}
\epsilon \leq \liminf _{k \rightarrow \infty}\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}} \tag{11}
\end{equation*}
$$

From (10) and (11), we get

$$
\begin{equation*}
\lim _{k \rightarrow \infty}\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}}=\epsilon \text { for any } l_{1}, l_{2} \in \mathbb{N} . \tag{12}
\end{equation*}
$$

We choose $l_{1}, l_{2} \in \mathbb{N}$ such that $\left(m_{k}+l_{2}\right)-\left(n_{k}+l_{1}\right)=1$.
From (7), we have

$$
\begin{gathered}
\psi\left(\left\|f \phi_{n_{k}+l_{1}}-f \phi_{m_{k}+l_{2}}\right\|_{E_{0}}\right) \leq \\
\psi\left(\left\|f \phi_{n_{k}+l_{1}-1}-f \phi_{m_{k}+l_{2}-1}\right\|_{E_{0}}\right)-\phi\left(| | f \phi_{n_{k}+l_{1}-1}-f \phi_{m_{k}+l_{2}-1} \|_{E_{0}}\right)
\end{gathered}
$$

On applying limits as $k \rightarrow \infty$ on both sides and by using (12), we get

$$
\psi(\epsilon) \leq \psi(\epsilon)-\eta(\epsilon)
$$

a contradiction.
Therefore, $\left\{f \phi_{n}\right\}$ is a Cauchy sequence in $f\left(R_{c}\right)$. Since $f\left(R_{c}\right)$ is complete, we have $f \phi_{n} \rightarrow \eta$ as $n \rightarrow \infty$ for some $\eta \in f\left(R_{c}\right)$ and hence there exists $\phi^{*} \in R_{c}$ such that $\eta=f \phi^{*}$ and $\lim _{n \rightarrow \infty} f \phi_{n}=f \phi^{*}$.
Now, for any $n \in \mathbb{N}$

$$
d\left(f \phi_{n+1}(c), T \phi^{*}\right) \leq H_{E}\left(T \phi_{n}, T \phi^{*}\right)
$$

and hence

$$
\begin{aligned}
\psi\left(d\left(f \phi_{n+1}(c), T \phi^{*}\right)\right) & \leq \psi\left(H_{E}\left(T \phi_{n}, T \phi^{*}\right)\right) \\
& \leq \psi\left(\left\|f \phi_{n}-f \phi^{*}\right\|_{E_{0}}\right)-\phi\left(\left\|f \phi_{n}-f \phi^{*}\right\|_{E_{0}}\right)
\end{aligned}
$$

On applying limits as $n \rightarrow \infty$ on both sides, we get
$\psi\left(d\left(f \phi^{*}(c), T \phi^{*}\right)\right) \leq \psi(0)-\phi(0)$ and hence $\psi\left(d\left(f \phi^{*}(c), T \phi^{*}\right)\right)=0$.
Therefore, $f \phi^{*}(c) \in T \phi^{*}$ and hence $T$ and $f$ have a PPF dependent coincidence point in $R_{c}$.

## 4. Existence of PPF Dependent Common Fixed Points

In this section, we introduce the concept of PPF dependent common fixed points for a pair of multi-valued mappings.

Definition 14. Let $S, T: E_{0} \rightarrow C B(E)$ be two multi-valued mappings. A point $\phi \in E_{0}$ is said to be a PPF dependent common fixed point of $S$ and $T$ if $\phi(c) \in S \phi$ and $\phi(c) \in T \phi$ for some $c \in I$.

In the following we define generalized weakly contractive type mappings for a pair of multi-valued mappings.

Definition 15. Let $S, T: E_{0} \rightarrow C B(E)$ be two multi-valued functions. The pair $(S, T)$ is said to be a pair of generalized weakly contractive type multi-valued mappings on $E_{0}$ if there exist $\psi \in \Psi$ and $\phi \in \Phi$ such that

$$
\begin{equation*}
\psi\left(H_{E}(T \alpha, S \beta)\right) \leq \psi(M(\alpha, \beta))-\phi(M(\alpha, \beta)) \tag{13}
\end{equation*}
$$

for any $\alpha, \beta \in E_{0}$, where
$M(\alpha, \beta)=\max \left\{\|\alpha-\beta\|_{E_{0}}, d(\alpha(c), T \alpha), d(\beta(c), S \beta), \frac{1}{2}[d(\beta(c), T \alpha)+d(\alpha(c), S \beta)]\right\}$.
Theorem 4. Let $S, T: E_{0} \rightarrow C B(E)$ be two multi-valued mappings such that:
(i) the pair $(S, T)$ is a pair of generalized weakly contractive type multi-valued mappings on $E_{0}$,
(ii) $R_{c}$ is algebraically closed with respect to the difference and
(iii) $T \phi \subseteq R_{c}(c)$ and $S \phi \subseteq R_{c}(c)$ for any $\phi \in E_{0}$.

Then, $S$ and $T$ have a PPF dependent common fixed point in $R_{c}$.
Proof. Let $\phi_{0} \in R_{c}$. Then, $T \phi_{0} \subseteq E$. Let $x_{1} \in E$ be such that $x_{1} \in T \phi_{0}$.
Since $T \phi_{0} \subseteq R_{c}(c)$, we choose $\phi_{1}$ in $R_{c}$ such that $x_{1}=\phi_{1}(c) \in T \phi_{0}$.
From (13), we have

$$
\psi\left(H_{E}\left(T \phi_{0}, S \phi_{1}\right)\right) \leq \psi\left(M\left(\phi_{0}, \phi_{1}\right)\right)-\phi\left(M\left(\phi_{0}, \phi_{1}\right)\right)
$$

If $M\left(\phi_{0}, \phi_{1}\right)=0$ then $\phi_{0}=\phi_{1}$ and hence $\phi_{0}$ is a PPF dependent common fixed point of $S$ and $T$.
Suppose that $M\left(\phi_{0}, \phi_{1}\right)>0$. By Lemma 1 there exists $x_{2} \in S \phi_{1}$ such that

$$
\begin{equation*}
\left\|x_{1}-x_{2}\right\|_{E} \leq H_{E}\left(T \phi_{0}, S \phi_{1}\right) \tag{14}
\end{equation*}
$$

Since $x_{2} \in S \phi_{1}$ and $S \phi_{1} \subseteq R_{c}(c)$, we choose $\phi_{2}$ in $R_{c}$ such that $x_{2}=\phi_{2}(c) \in S \phi_{1}$. From (13), we have

$$
\psi\left(H_{E}\left(S \phi_{1}, T \phi_{2}\right)\right)=\psi\left(H_{E}\left(T \phi_{2}, S \phi_{1}\right)\right) \leq \psi\left(M\left(\phi_{2}, \phi_{1}\right)\right)-\phi\left(M\left(\phi_{2}, \phi_{1}\right)\right)
$$

If $M\left(\phi_{2}, \phi_{1}\right)=0$ then $\phi_{1}=\phi_{2}$ and hence $\phi_{1}$ is a PPF dependent common fixed point of $S$ and $T$.
Suppose that $M\left(\phi_{2}, \phi_{1}\right)>0$. By Lemma 1 there exists $x_{3} \in T \phi_{2}$ such that

$$
\begin{equation*}
\left\|x_{2}-x_{3}\right\|_{E} \leq H_{E}\left(S \phi_{1}, T \phi_{2}\right) \tag{15}
\end{equation*}
$$

Since $x_{3} \in T \phi_{2}$ and $T \phi_{2} \subseteq R_{c}(c)$, we choose $\phi_{3}$ in $R_{c}$ such that $x_{3}=\phi_{3}(c) \in T \phi_{2}$. Again from (13), we have

$$
\psi\left(H_{E}\left(T \phi_{2}, S \phi_{3}\right)\right) \leq \psi\left(M\left(\phi_{2}, \phi_{3}\right)\right)-\phi\left(M\left(\phi_{2}, \phi_{3}\right)\right)
$$

If $M\left(\phi_{2}, \phi_{3}\right)=0$ then $\phi_{2}=\phi_{3}$ and hence $\phi_{2}$ is a PPF dependent common fixed point of $S$ and $T$.
Suppose that $M\left(\phi_{2}, \phi_{3}\right)>0$. On continuing this process, we get a sequence $\left\{\phi_{n}\right\}$ in $R_{c}$ such that

$$
\begin{equation*}
\phi_{2 n+1}(c) \in T \phi_{2 n}, \quad \phi_{2 n+2}(c) \in S \phi_{2 n+1} \tag{16}
\end{equation*}
$$

and

$$
\begin{equation*}
M\left(\phi_{n}, \phi_{n+1}\right)>0 \tag{17}
\end{equation*}
$$

with $\left\|\phi_{2 n+1}(c)-\phi_{2 n+2}(c)\right\|_{E} \leq H_{E}\left(T \phi_{2 n}, S \phi_{2 n+1}\right)$
and $\left\|\phi_{2 n+2}(c)-\phi_{2 n+3}(c)\right\|_{E} \leq H_{E}\left(S \phi_{2 n+1}, T \phi_{2 n+2}\right)$ for all $n \in \mathbb{N} \cup\{0\}$.
Since $R_{c}$ is algebraically closed with respect to the difference, for all $n \in \mathbb{N} \cup\{0\}$ we have

$$
\begin{equation*}
\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}} \leq H_{E}\left(T \phi_{2 n}, S \phi_{2 n+1}\right) \tag{18}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|\phi_{2 n+2}-\phi_{2 n+3}\right\|_{E_{0}} \leq H_{E}\left(S \phi_{2 n+1}, T \phi_{2 n+2}\right)=H_{E}\left(T \phi_{2 n+2}, S \phi_{2 n+1}\right) \tag{19}
\end{equation*}
$$

We consider

$$
\begin{aligned}
& M\left(\phi_{2 n}, \phi_{2 n+1}\right)=\max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}}, d\left(\phi_{2 n}(c), T \phi_{2 n}\right), d\left(\phi_{2 n+1}(c), S \phi_{2 n+1}\right),\right. \\
&\left.\frac{1}{2}\left[d\left(\phi_{2 n+1}(c), T \phi_{2 n}\right)+d\left(\phi_{2 n}(c), S \phi_{2 n+1}\right)\right]\right\}, \\
& \leq \max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n}(c)-\phi_{2 n+1}(c)\right\|_{E},\left\|\phi_{2 n+1}(c)-\phi_{2 n+2}(c)\right\|_{E},\right. \\
& \frac{1}{2}\left[0+\left\|\phi_{2 n}(c)-\phi_{2 n+2}(c)\right\|_{E}\right\} \\
&= \max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}, \frac{1}{2}\left[\left\|\phi_{2 n}-\phi_{2 n+2}\right\|_{E_{0}}\right]\right\} \\
& \leq \max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right. \\
&\left.\frac{1}{2}\left[\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}}+\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right]\right\} \\
&= \max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right\},
\end{aligned}
$$

and hence

$$
\begin{equation*}
M\left(\phi_{2 n}, \phi_{2 n+1}\right) \leq \max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right\} \tag{20}
\end{equation*}
$$

Suppose that $\max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right\}=\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}$. Now, from 20), we have

$$
M\left(\phi_{2 n}, \phi_{2 n+1}\right) \leq\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}
$$

and hence

$$
\psi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right) \leq \psi\left(\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right)
$$

Now, from (18), we have

$$
\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}} \leq H_{E}\left(T \phi_{2 n}, S \phi_{2 n+1}\right)
$$

and hence

$$
\begin{align*}
\psi\left(\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right) & \leq \psi\left(H_{E}\left(T \phi_{2 n}, S \phi_{2 n+1}\right)\right) \\
& \leq \psi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)-\phi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)  \tag{21}\\
& \leq \psi\left(\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right)-\phi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)
\end{align*}
$$

Therefore, $f\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)=0$ and hence $M\left(\phi_{2 n}, \phi_{2 n+1}\right)=0$, a contradiction.
Therefore,

$$
\begin{equation*}
\max \left\{\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}},\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right\}=\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}} \tag{22}
\end{equation*}
$$

Now, from 20, we have

$$
\begin{equation*}
M\left(\phi_{2 n}, \phi_{2 n+1}\right) \leq\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}} \tag{23}
\end{equation*}
$$

Now, from (18), we have

$$
\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}} \leq H_{E}\left(T \phi_{2 n}, S \phi_{2 n+1}\right)
$$

and hence

$$
\begin{aligned}
\psi\left(\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right) & \leq \psi\left(H_{E}\left(T \phi_{2 n}, S \phi_{2 n+1}\right)\right) \\
& \leq \psi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)-\phi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right) \\
& <\psi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)(\text { by using (17) } \\
& \leq \psi\left(\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}}\right) .(\text { by using 23) })
\end{aligned}
$$

Since $\psi$ is monotonically increasing function, we have

$$
\begin{equation*}
\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}} \leq M\left(\phi_{2 n}, \phi_{2 n+1}\right) \leq\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}} \tag{24}
\end{equation*}
$$

Similarly we have $\left\|\phi_{2 n+2}-\phi_{2 n+3}\right\|_{E_{0}} \leq M\left(\phi_{2 n+2}, \phi_{2 n+1}\right) \leq\left\|\phi_{2 n+2}-\phi_{2 n+1}\right\|_{E_{0}}$

$$
\begin{equation*}
=\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}} \tag{25}
\end{equation*}
$$

From (24) and (25), we have $\left\|\phi_{n+1}-\phi_{n}\right\|_{E_{0}} \leq\left\|\phi_{n}-\phi_{n-1}\right\|_{E_{0}}$ for all $n \in \mathbb{N}$.
Therefore, the sequence $\left\{\left\|\phi_{n+1}-\phi_{n}\right\|_{E_{0}}\right\}$ is a decreasing sequence in $\mathbb{R}^{+}$, and hence convergent.
Let $\lim _{n \rightarrow 1}\left\|\phi_{n+1}-\phi_{n}\right\|_{E_{0}}=r$ (say).
From (24), we have

$$
\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}} \leq M\left(\phi_{2 n}, \phi_{2 n+1}\right) \leq\left\|\phi_{2 n}-\phi_{2 n+1}\right\|_{E_{0}}
$$

On applying limits as $n \rightarrow \infty$, we get

$$
r \leq \lim _{n \rightarrow \infty} M\left(\phi_{2 n}, \phi_{2 n+1}\right) \leq r \text { and hence } \lim _{n \rightarrow \infty} M\left(\phi_{2 n}, \phi_{2 n+1}\right)=r
$$

From (21), we have

$$
\psi\left(\left\|\phi_{2 n+1}-\phi_{2 n+2}\right\|_{E_{0}}\right) \leq \psi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)-\phi\left(M\left(\phi_{2 n}, \phi_{2 n+1}\right)\right)
$$

On applying limits as $n \rightarrow \infty$, we get $\psi(r) \leq \psi(r)-\phi(r)$ and which implies that $r=0$.
Therefore,

$$
\begin{equation*}
\lim _{n \rightarrow \infty}\left\|\phi_{n+1}-\phi_{n}\right\|_{E_{0}}=0 \tag{26}
\end{equation*}
$$

Now, we show that $\left\{\phi_{n}\right\}$ is a Cauchy sequence.
From (26), to prove $\left\{\phi_{n}\right\}$ is a Cauchy sequence it is enough to prove that $\left\{\phi_{2 n}\right\}$ is a Cauchy sequence.
Suppose that $\left\{\phi_{2 n}\right\}$ is not a Cauchy sequence.
Then, there exists $\epsilon>0$ and two subsequences $\left\{\phi_{2 m_{k}}\right\}$ and $\left\{\phi_{2 n_{k}}\right\}$ of $\left\{\phi_{2 n}\right\}$ such that for any $k \in \mathbb{N}, m_{k}>n_{k}>k$ such that

$$
\begin{equation*}
\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}}\right\|_{E_{0}} \geq \epsilon \tag{27}
\end{equation*}
$$

Let $m_{k}$ be the smallest positive integer greater than $n_{k}$ that is satisfying (27).
Then, $\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}}\right\|_{E_{0}} \geq \epsilon$ and $\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}-2}\right\|_{E_{0}}<\epsilon$.
We now show that $\lim _{k \rightarrow \infty}\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}}=\epsilon$.
Clearly

$$
\epsilon \leq\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}}\right\|_{E_{0}} \leq\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}}+\left\|\phi_{2 m_{k}+1}-\phi_{2 m_{k}}\right\|_{E_{0}} .
$$

Now, by applying Proposition 1 with $a_{k}=\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}}$ and
$b_{k}=\left\|\phi_{2 m_{k}+1}-\phi_{2 m_{k}}\right\|_{E_{0}}$ we have

$$
\epsilon \leq \liminf _{k \rightarrow \infty}\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}}+\limsup _{k \rightarrow \infty}\left\|\phi_{2 m_{k}+1}-\phi_{2 m_{k}}\right\|_{E_{0}}
$$

and hence

$$
\begin{equation*}
\epsilon \leq \liminf _{k \rightarrow \infty}\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}} . \tag{28}
\end{equation*}
$$

Clearly

$$
\begin{aligned}
\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}} & \leq\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}-2}\right\|_{E_{0}}+\left\|\phi_{2 m_{k}-2}-\phi_{2 m_{k}-1}\right\|_{E_{0}} \\
& +\left\|\phi_{2 m_{k}-1}-\phi_{2 m_{k}}\right\|_{E_{0}}+\left\|\phi_{2 m_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}} \\
& <\epsilon+\left\|\phi_{2 m_{k}-2}-\phi_{2 m_{k}-1}\right\|_{E_{0}}
\end{aligned}
$$

$+\left\|\phi_{2 m_{k}-1}-\phi_{2 m_{k}}\right\|_{E_{0}}+\left\|\phi_{2 m_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}}$.
On applying limit superior as $k \rightarrow \infty$ on both sides, we get

$$
\begin{equation*}
\limsup _{k \rightarrow \infty}\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}} \leq \epsilon . \tag{29}
\end{equation*}
$$

From (28) and (29), we get

$$
\begin{equation*}
\lim _{k \rightarrow \infty}\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}}=\epsilon \tag{30}
\end{equation*}
$$

We now show that $\lim _{k \rightarrow \infty}\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}}=\epsilon$ for any $l_{1}, l_{2} \in \mathbb{N}$.
Let $l_{1}, l_{2} \in \mathbb{N}$.
We now consider

$$
\begin{aligned}
\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}} & \leq\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 n_{k}+l_{1}-1}\right\|_{E_{0}}+\left\|\phi_{2 n_{k}+l_{1}-1}-\phi_{2 n_{k}+l_{2}-2}\right\|_{E_{0}} \\
& +\ldots+\left\|\phi_{2 n_{k}+1}-\phi_{2 n_{k}}\right\|_{E_{0}}+\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}} \\
& +\left\|\phi_{2 m_{k}+1}-\phi_{2 m_{k}+2}\right\|_{E_{0}}+\ldots+\left\|\phi_{2 m_{k}+l_{2}-1}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}} .
\end{aligned}
$$

On applying limit superior as $k \rightarrow \infty$ on both sides, we get

$$
\begin{equation*}
\limsup _{k \rightarrow \infty}\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}} \leq \epsilon \tag{31}
\end{equation*}
$$

We now consider

$$
\begin{aligned}
\left\|\phi_{2 n_{k}}-\phi_{2 m_{k}+1}\right\|_{E_{0}} & \leq\left\|\phi_{2 n_{k}}-\phi_{2 n_{k}+1}\right\|_{E_{0}}+\left\|\phi_{2 n_{k}+1}-\phi_{2 n_{k}+2}\right\|_{E_{0}}+\ldots \\
& +\left\|\phi_{2 n_{k}+l_{1}-1}-\phi_{2 n_{k}+l_{1}}\right\|_{E_{0}}+\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}} \\
& +\left\|\phi_{2 m_{k}+l_{2}}-\phi_{2 m_{k}+l_{2}-1}\right\|_{E_{0}}+\ldots+\left\|\phi_{2 m_{k}+2}-\phi_{2 m_{k}+1}\right\|_{E_{0}} .
\end{aligned}
$$

Now, by applying Proposition 1 with $a_{k}=\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}}$ and
$b_{k}=\left(\left\|\phi_{2 n_{k}}-\phi_{2 n_{k}+1}\right\|_{E_{0}}+\left\|\phi_{2 n_{k}+1}-\phi_{2 n_{k}+2}\right\|_{E_{0}}+\ldots+\left\|\phi_{2 n_{k}+l_{1}-1}-\phi_{2 n_{k}+l_{1}}\right\|_{E_{0}}+\right.$ $\left.\left\|\phi_{2 m_{k}+l_{2}}-\phi_{2 m_{k}+l_{2}-1}\right\|_{E_{0}}+\ldots+\left\|\phi_{2 m_{k}+2}-\phi_{2 m_{k}+1}\right\|_{E_{0}}\right)$
we have

$$
\begin{aligned}
\epsilon & \leq \liminf _{k \rightarrow \infty}\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}}+\limsup _{k \rightarrow \infty}\left(\left\|\phi_{2 n_{k}}-\phi_{2 n_{k}+1}\right\|_{E_{0}}\right. \\
& +\left\|\phi_{2 n_{k}+1}-\phi_{2 n_{k}+2}\right\|_{E_{0}}+\ldots+\left\|\phi_{2 n_{k}+l_{1}-1}-\phi_{2 n_{k}+l_{1}}\right\|_{E_{0}} \\
& \left.+\left\|\phi_{2 m_{k}+l_{2}}-\phi_{2 m_{k}+l_{2}-1}\right\|_{E_{0}}+\ldots+\left\|\phi_{2 m_{k}+2}-\phi_{2 m_{k}+1}\right\|_{E_{0}}\right) .
\end{aligned}
$$

Hence

$$
\begin{equation*}
\epsilon \leq \liminf _{k \rightarrow \infty}\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}} . \tag{32}
\end{equation*}
$$

From (31) and (32), we get that for any $l_{1}, l_{2} \in \mathbb{N}$

$$
\begin{equation*}
\lim _{k \rightarrow \infty}\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}}=\epsilon . \tag{33}
\end{equation*}
$$

Now, we choose $l_{1}, l_{2} \in \mathbb{N}$ such that $2 n_{k}+l_{1}$ is even, $2 m_{k}+l_{2}$ is odd and $\left(2 m_{k}+l_{2}\right)-\left(2 n_{k}+l_{1}\right)=1$.
From (24), we have

$$
\left\|\phi_{2 n_{k}+l_{1}+1}-\phi_{2 m_{k}+l_{2}+1}\right\|_{E_{0}} \leq M\left(\phi_{2 n_{k}+l_{1}}, \phi_{2 m_{k}+l_{2}}\right) \leq\left\|\phi_{2 n_{k}+l_{1}}-\phi_{2 m_{k}+l_{2}}\right\|_{E_{0}} .
$$

On applying limits as $k \rightarrow \infty$, we get
$\epsilon \leq \lim _{k \rightarrow \infty} M\left(\phi_{2 n_{k}+l_{1}}, \phi_{2 m_{k}+l_{2}}\right) \leq \epsilon$ and hence $\lim _{k \rightarrow \infty} M\left(\phi_{2 n_{k}+l_{1}}, \phi_{2 m_{k}+l_{2}}\right)=\epsilon$.
From (21), we have

$$
\left.\overline{\psi(\|} \mid \phi_{2 n_{k}+l_{1}+1}-\phi_{2 m_{k}+l_{2}+1} \|_{E_{0}}\right) \leq \psi\left(M\left(\phi_{2 n_{k}+l_{1}}, \phi_{2 m_{k}+l_{2}}\right)\right)-\phi\left(M\left(\phi_{2 n_{k}+l_{1}}, \phi_{2 m_{k}+l_{2}}\right)\right)
$$

On applying limits as $k \rightarrow \infty$ we get,

$$
\psi(\epsilon) \leq \psi(\epsilon)-\phi(\epsilon) \text { and hence } \epsilon=0
$$

a contradiction.
Therefore, the sequence $\left\{\phi_{n}\right\}$ is a Cauchy sequence in $R_{c}$.
Since $E_{0}$ is complete, we have $\phi_{n} \rightarrow \phi^{*}$ as $n \rightarrow \infty$ for some $\phi^{*} \in E_{0}$.
Since $R_{c}$ is topologically closed, we have $\phi^{*} \in R_{c}$.
Now, we show that $\phi^{*}$ is a PPF dependent common fixed point of $S$ and $T$.
We now consider,
$d\left(\phi^{*}(c), S \phi^{*}\right) \leq M\left(\phi_{2 k}, \phi^{*}\right)$

$$
\begin{aligned}
& =\max \left\{\left\|\phi_{2 k}-\phi^{*}\right\|_{E_{0}}, d\left(\phi_{2 k}(c), T \phi_{2 k}\right), d\left(\phi^{*}(c), S \phi^{*}\right),\right. \\
& \left.\quad \frac{1}{2}\left[d\left(\phi^{*}(c), T \phi_{2 k}\right)+d\left(\phi_{2 k}(c), S \phi^{*}\right)\right]\right\} \\
& \leq \max \left\{\left\|\phi_{2 k}-\phi^{*}\right\|_{E_{0}},\left\|\phi_{2 k}(c)-\phi_{2 k+1}(c)\right\|_{E}+d\left(\phi_{2 k+1}(c), T \phi_{2 k}\right), d\left(\phi^{*}(c), S \phi^{*}\right),\right. \\
& \quad \frac{1}{2}\left[\left\|\phi^{*}(c)-\phi_{2 k+1}(c)\right\|_{E}+d\left(\phi_{2 k+1}(c), T \phi_{2 k}\right)\right. \\
& \left.\left.\quad+\left\|\phi_{2 k}(c)-\phi^{*}(c)\right\|_{E}+d\left(\phi^{*}(c), S \phi^{*}\right)\right]\right\} \\
& =\max \left\{\left\|\phi_{2 k}-\phi^{*}\right\|_{E_{0}},\left\|\phi_{2 k}-\phi_{2 k+1}\right\|_{E_{0}}, d\left(\phi^{*}(c), S \phi^{*}\right),\right. \\
& \left.\quad \frac{1}{2}\left[\left\|\phi^{*}-\phi_{2 k+1}\right\|_{E_{0}}+\left\|\phi_{2 k}-\phi^{*}\right\|_{E_{0}}+d\left(\phi^{*}(c), S \phi^{*}\right)\right]\right\} .
\end{aligned}
$$

On applying limits as $k \rightarrow \infty$, we get

$$
\begin{aligned}
d\left(\phi^{*}(c), S \phi^{*}\right) & \leq \lim _{k \rightarrow \infty} M\left(\phi_{2 k}, \phi^{*}\right) \\
& \leq \max \left\{0,0, d\left(\phi^{*}(c), S \phi^{*}\right), \frac{1}{2}\left[d\left(\phi^{*}(c), S \phi^{*}\right)\right]\right\} \\
& =d\left(\phi^{*}(c), S \phi^{*}\right)
\end{aligned}
$$

Hence $\lim _{k \rightarrow \infty} M\left(\phi_{2 k}, \phi^{*}\right)=d\left(\phi^{*}(c), S \phi^{*}\right)$.
Now,

$$
\begin{aligned}
\left.d\left(\phi^{*}(c), S \phi^{*}\right)\right) & \leq\left\|\phi^{*}(c)-\phi_{2 k+1}(c)\right\|_{E}+d\left(\phi_{2 k+1}(c), S \phi^{*}\right) \\
& \leq\left\|\phi^{*}-\phi_{2 k+1}\right\|_{E_{0}}+H_{E}\left(T \phi_{2 k}, S \phi^{*}\right)
\end{aligned}
$$

Applying limits as $k \rightarrow \infty$, we get

$$
\left.d\left(\phi^{*}(c), S \phi^{*}\right)\right) \leq \lim _{k \rightarrow \infty} H_{E}\left(T \phi_{2 k}, S \phi^{*}\right)
$$

and hence

$$
\begin{aligned}
\psi\left(d\left(\phi^{*}(c), S \phi^{*}\right)\right) & \leq \lim _{k \rightarrow \infty} \psi\left(H_{E}\left(T \phi_{2 k}, S \phi^{*}\right)\right) \\
& \leq \lim _{k \rightarrow \infty} \psi\left(M\left(\phi_{2 k}, \phi^{*}\right)\right)-\lim _{k \rightarrow \infty} \phi\left(M\left(\phi_{2 k}, \phi^{*}\right)\right) \\
& =\psi\left(d\left(\phi^{*}(c), S \phi^{*}\right)\right)-\phi\left(d\left(\phi^{*}(c), S \phi^{*}\right)\right)
\end{aligned}
$$

Therefore, $\phi\left(d\left(\phi^{*}(c), S \phi^{*}\right)\right)=0$ and hence $\phi^{*}(c) \in S \phi^{*}$.
Similarly we can prove that $\phi^{*}(c) \in T \phi^{*}$.
Therefore, $\phi^{*}$ is a PPF dependent common fixed point of $S$ and $T$.

## 5. Corollaries and Examples

Corollary 1. Let $T: E_{0} \rightarrow C B(E)$ and $f: E \rightarrow E$ be a function that satisfy the following conditions:
(i) $T$ is weakly contractive type multi-valued mapping with respect to $f$,
(ii) $T \phi \subseteq f\left(R_{c}\right)(c)$ for any $\phi \in E_{0}$,
(iii) $R_{c}$ is algebraically closed with respect to the difference,
(iv) $f\left(R_{c}\right)$ is complete and
(v) $f\left(R_{c}\right) \subseteq R_{c}$.

Then, $T$ and $f$ have a PPF dependent coincidence point in $R_{c}$.
Proof. Follows from Theorem 3 by choosing $\psi(t)=t, t \in \mathbb{R}^{+}$in the inequality (1).

By choosing $f=I, I$ the identity map in Theorem 3, we get the following corollary.
Corollary 2. Let $T: E_{0} \rightarrow C B(E)$ be a multi-valued mapping. Assume that $T$ satisfy the following conditions:
(i) $T$ is a generalized weakly contractive type multi-valued mapping,
(ii) $T \phi \subseteq R_{c}(c)$ for any $\phi \in E_{0}$,
(iii) $R_{c}$ is algebraically closed with respect to the difference.

Then, $T$ has a PPF dependent fixed point in $R_{c}$.
The following corollary follows by choosing $\psi(t)=t, t \in \mathbb{R}^{+}$in Corollary 2 .
Corollary 3. Let $T: E_{0} \rightarrow C B(E)$ be a mapping satisfy the following conditions:
(i) $T$ is weakly contractive type multi-valued mapping,
(ii) $T \phi \subseteq R_{c}(c)$ for any $\phi \in E_{0}$,
(iii) $R_{c}$ is algebraically closed with respect to the difference.

Then, $T$ has a PPF dependent fixed point in $R_{c}$.
Corollary 4. Let $T: E_{0} \rightarrow C B(E)$ be a mapping satisfiy the following conditions:
(i) suppose that there exists $k \in[0,1)$ such that

$$
\left.H_{E}(T \alpha, T \beta)\right) \leq k\|\alpha-\beta\|_{E_{0}} \text { for all } \alpha, \beta \in E_{0}
$$

(ii) $T \phi \subseteq R_{c}(c)$ for any $\phi \in E_{0}$,
(iii) $R_{c}$ is algebraically closed with respect to the difference.

Then, $T$ has a PPF dependent fixed point in $R_{c}$.
Proof. Follows by choosing $\phi(t)=(1-k) t, t \in \mathbb{R}^{+}$in Corollary 3 .
Corollary 5. Let $S, T: E_{0} \rightarrow C B(E)$ be two multi-valued mappings such that
(i) $\left.H_{E}(T \alpha, S \beta)\right) \leq k \max \left\{\|\alpha-\beta\|_{E_{0}}, d(\alpha(c), T \alpha), d(\beta(c), S \beta), \frac{1}{2}[d(\beta(c), T \alpha)+\right.$ $d(\alpha(c), S \beta)]\}$ for any $\alpha, \beta \in E_{0}$,
(ii) $R_{c}$ is algebraically closed with respect to the difference and
(iii) $T \alpha \subseteq R_{c}(c)$ and $S \alpha \subseteq R_{c}(c)$ for all $\alpha \in E_{0}$.

Then, $S$ and $T$ have a PPF dependent common fixed point in $R_{c}$.
Proof. Follows by choosing $\psi(t)=t$ and $\phi(t)=(1-k) t$ for $t \in \mathbb{R}^{+}$in Theorem 4.

If $S=T$ in Theorem 4 and Corollary 5, we get the following corollaries.

Corollary 6. Let $T: E_{0} \rightarrow C B(E)$ be a multi-valued mapping. Assume that:
(i) there exist two functions $\psi \in \Psi$ and $\phi \in \Phi$ such that

$$
\begin{equation*}
\psi\left(H_{E}(T \alpha, T \beta)\right) \leq \psi(M(\alpha, \beta))-\phi(M(\alpha, \beta)) \tag{34}
\end{equation*}
$$

for all $\alpha, \beta \in E_{0}$, where
$M(\alpha, \beta)=\max \left\{\|\alpha-\beta\|_{E_{0}}, d(\alpha(c), T \alpha), d(\beta(c), T \beta), \frac{1}{2}[d(\beta(c), T \alpha)+d(\alpha(c), T \beta)]\right\}$,
(ii) $R_{c}$ is algebraically closed with respect to the difference and
(iii) $T \phi \subseteq R_{c}(c)$ for any $\phi \in E_{0}$.

Then, $T$ has a PPF dependent fixed point in $R_{c}$.
Corollary 7. Let $T: E_{0} \rightarrow C B(E)$ be two multi-valued mappings such that
(i) $\left.H_{E}(T \alpha, T \beta)\right) \leq k \max \left\{\|\alpha-\beta\|_{E_{0}}, d(\alpha(c), T \alpha), d(\beta(c), T \beta), \frac{1}{2}[d(\beta(c), T \alpha)+\right.$ $d(\alpha(c), T \beta)]\}$
for all $\alpha, \beta \in E_{0}$,
(ii) $R_{c}$ is algebraically closed with respect to the difference and
(iii) $T \alpha \subseteq R_{c}(c)$ for any $\alpha \in E_{0}$.

Then, $T$ has a PPF dependent fixed point in $R_{c}$.
Example 1. Let $E=\mathbb{R}, c=1 \in I=\left[\frac{1}{2}, 2\right] \subseteq \mathbb{R}, E_{0}=C(I, E)$.
Let $k \geq 1$. We define $f: E \rightarrow E$ by $f(x)=k x$ for any $x \in E$.
Clearly, $f$ is a continuous function.
By definition, $R_{c}(c)=\left\{\phi(c) \mid \phi \in R_{c}\right\}$ and

$$
f\left(R_{c}\right)(c)=\left\{(f \circ \phi)(c) \mid \phi \in R_{c}\right\}=\left\{f(\phi(c)) \mid \phi \in R_{c}\right\}=\left\{k \phi(c) \mid \phi \in R_{c}\right\}
$$

First we show that $f\left(R_{c}\right)=R_{c}$.
Let $\alpha \in R_{c}$. Then $\alpha=\beta$ for some $\beta \in R_{c}$.
Clearly, $\alpha=k \frac{1}{k} \beta=k \eta \quad$ (by Theorem 1, $\eta=\frac{1}{k} \beta \in R_{c}$ ) so that
$\alpha(x)=k \eta(x)=f(\eta(x))=(f \circ \eta)(x)$ for any $x \in I$.
Therefore, $\alpha=f \circ \eta \in f\left(R_{c}\right)$ and hence

$$
\begin{equation*}
R_{c} \subseteq f\left(R_{c}\right) \tag{35}
\end{equation*}
$$

Now, let $\alpha \in f\left(R_{c}\right)$. Then $\alpha=f \circ \beta$ for some $\beta \in R_{c}$.
Clearly, $\alpha(x)=(f \circ \beta)(x)=f(\beta(x))=k \beta(x)=(k \beta)(x)$ for any $x \in I$.
Therefore, $\alpha=k \beta \in R_{c}$ and hence

$$
\begin{equation*}
f\left(R_{c}\right) \subseteq R_{c} \tag{36}
\end{equation*}
$$

From (35) and (36), we get $f\left(R_{c}\right)=R_{c}$.
Since $E_{0}$ is complete and $R_{c}$ is topologically closed we have $f\left(R_{c}\right)=R_{c}$ is complete.
For any $\gamma \in \mathbb{R}$, we define $\phi_{\gamma}: I \rightarrow E$ by

$$
\phi_{\gamma}(x)= \begin{cases}\gamma x^{2} & \text { if } x \in\left[\frac{1}{2}, 1\right] \\ \frac{\gamma}{x^{2}} & \text { if } x \in[1,2]\end{cases}
$$

Clearly $\phi_{\gamma} \in E_{0},\left\|\phi_{\gamma}\right\|_{E_{0}}=\left\|\phi_{\gamma}(c)\right\|_{E}$ and hence $\phi_{\gamma} \in R_{c}$ for any $\gamma \in \mathbb{R}$.
Let $\mathrm{F}_{0}=\left\{\phi_{\gamma} \mid \gamma \in \mathbb{R}\right\}$.
Then, $\mathrm{F}_{0}$ is algebraically closed with respect to the difference and $\mathrm{F}_{0} \subseteq R_{c}$.

We observe that $\mathbb{R}=\left\{\phi_{\gamma}(c) \mid \gamma \in \mathbb{R}\right\}=\mathrm{F}_{0}(c) \subseteq R_{c}(c)$.
Clearly, $R_{c}(c) \subseteq \mathbb{R}$ and hence $f\left(R_{c}\right)(c)=R_{c}(c)=\mathbb{R}$.
We define $T: E_{0} \rightarrow C B(E)$ by $T \phi=\left[0, \frac{k}{4}\|\phi(c)\|_{E}\right]$ for any $\phi \in E_{0}$.
Clearly, $T \phi \subseteq \mathbb{R}=R_{c}(c)=f\left(R_{c}\right)(c)$.
We define $\psi, \phi: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$by $\psi(t)=\frac{t^{2}}{2}$ and

$$
\phi(t)= \begin{cases}\frac{15 t^{3}}{32} & \text { if } t \in[0,1] \\ \frac{15^{t} t}{32} & \text { if } t \geq 1\end{cases}
$$

Clearly, $\psi \in \Psi$ and $\phi \in \Phi$.
From the definition of Hausdorff distance, it follows that, for any $\alpha, \beta \in E_{0}$

$$
\begin{aligned}
H_{E}(T \alpha, T \beta)=\frac{k}{4}\left\{\begin{array}{l}
\|\alpha(c)\|_{E}-\|\beta(c)\|_{E} \quad \text { if }\|\alpha(c)\|_{E} \geq\|\beta(c)\|_{E} \\
\|\beta(c)\|_{E}-\|\alpha(c)\|_{E} \quad \text { if }\|\beta(c)\|_{E} \geq\|\alpha(c)\|_{E}
\end{array}\right. \\
\quad \begin{array}{l}
\quad=\frac{k}{4}\left|\|\alpha(c)\|_{E}-\|\beta(c)\|_{E}\right|=\frac{1}{4}\left|\|k \alpha(c)\|_{E}-\|k \beta(c)\|_{E}\right| \\
\\
\leq \frac{1}{4}|k \alpha(c)-k \beta(c)|=\frac{1}{4}|(f \circ \alpha)(c)-(f \circ \beta)(c)| \\
\\
\quad=\frac{1}{4}\|(f \alpha-f \beta)(c)\|_{E} \\
\\
\leq \frac{1}{4}\|f \alpha-f \beta\|_{E_{0}} .
\end{array}
\end{aligned}
$$

Therefore,

$$
\begin{aligned}
& \psi\left(H_{E}(T \alpha, T \beta)\right) \leq \psi\left(\frac{1}{4}\|f \alpha-f \beta\|_{E_{0}}\right)=\frac{1}{32}\left[\|f \alpha-f \beta\|_{E_{0}}\right]^{2} \\
& \leq \psi\left(\|f \alpha-f \beta\| \|_{E_{0}}\right)-\phi\left(\|f \alpha-f \beta\| \|_{E_{0}}\right)
\end{aligned}
$$

Therefore, $T$ and $f$ satisfy all the hypotheses of Theorem 3 and $\phi_{0} \in R_{c}$ is a PPF dependent coincidence point of $T$ and $f$.

Example 2. Let $E=\mathbb{R}, c=1 \in I=\left[\frac{1}{2}, 2\right] \subseteq \mathbb{R}, E_{0}=C(I, E)$.
On continuing the same procedure as in the Example 1, we get $R_{c}(c)=\mathbb{R}$.
We define $T: E_{0} \rightarrow C B(E)$ by $T \phi=\left[0, \frac{1}{5}\|\phi(c)\|_{E}\right]$ for any $\phi \in E_{0}$.
Clearly $T \phi \subseteq R_{c}(c)$.
We define $\psi, \phi: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$by $\psi(t)=2 t$ and $\phi(t)=\frac{6 t}{5}$ for any $t \in \mathbb{R}^{+}$.
Clearly, $\psi \in \Psi$ and $\phi \in \Phi$.
Clearly, for any $\alpha, \beta \in E_{0}$, we have

$$
\begin{aligned}
H_{E}(T \alpha, T \beta) & \leq \frac{1}{5}\|\alpha-\beta\|_{E_{0}} \\
& \leq \frac{1}{5} \max \left\{\|\alpha-\beta\|_{E_{0}}, d(\alpha(c), T \alpha), d(\beta(c), T \beta), \frac{1}{2}[d(\beta(c), T \alpha)+\right. \\
& =\frac{1}{5} M(\alpha, \beta)
\end{aligned}
$$

$d(\alpha(c), T \beta)]\}$
Therefore,

$$
\begin{aligned}
\psi\left(H_{E}(T \alpha, T \beta)\right) & \leq \psi\left(\frac{1}{5} M(\alpha, \beta)\right)=\frac{2}{5} M(\alpha, \beta) \\
& \leq 2 M(\alpha, \beta)-\frac{6}{5} M(\alpha, \beta) \\
& =\psi(M(\alpha, \beta))-\phi(M(\alpha, \beta))
\end{aligned}
$$

Therefore, $T$ satisfies all the hypotheses of Corollary 6 and $\phi_{0} \in R_{c}$ is a PPF dependent fixed point of $T$.

Author Contribution Statements Each author declares substantial contributions through the following: (1) the conception and design of the study, or acquisition of data, or analysis and interpretation of data, (2) drafting the article or revising it critically for important intellectual content.

Declaration of Competing Interests The authors declare that they have no known competing financial interests or personal relationships that could have appeared to influence the work reported in this paper.

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# MAJORIZATION PROPERTY FOR CERTAIN CLASSES OF ANALYTIC FUNCTIONS ASSOCIATED WITH GENERAL OPERATOR 

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Abstract. In this study, we introduce two new classes $S_{k}[E, F ; \mu ; \gamma]$ and $T_{k}(\theta, \mu, \gamma)$ of analytic functions using the general integral operator. For these two classes, we study the majorization properties. Some applications of the results are discussed in the form of corollaries.

## 1. Introduction and Definitions

The Majorization for two analytic functions $u$ and $v$ is defined as follows (see 17)

$$
u(\xi) \prec \prec v(\xi) ; \quad(\xi \in D),
$$

if there is an analytic function $\psi(\xi)$, such that

$$
\begin{equation*}
|\psi(\xi)| \leq 1 \text { and } u(\xi)=\psi(\xi) v(\xi) ; \quad(\xi \in D) \tag{1}
\end{equation*}
$$

where $D=\{\xi \in \mathbb{C}:|\xi|<1\}$ is an open unit disk.
The function $u$ is subordinate to $v$ and defined as $u(\xi) \prec v(\xi)$, if there is a schwarz function $w$, that is analytic in $D$ with $|w(\xi)|<1, w(0)=0, \xi \in D$ such that $u(\xi)=v(w(\xi)), \xi \in D$.

Thus, by combining subordination and majorization, we may define quasi-subordination as follows:

[^8]We say that the function $u$ is quasi-subordinate relative to $\phi(z)$ to the function $v$ and defined as (See 19)

$$
u(\xi) \prec_{q} v(\xi) ; \quad(\xi \in D)
$$

If there are two analytic functions $\psi(\xi)$ and $w(\xi)$ in $D$ such that $\frac{u(\xi)}{\psi(\xi)}$ is analytic and subordinate to $v(\xi)$ in $D$ and

$$
|\psi(\xi)| \leq 1 \text { and } w(0)=0, \quad|w(\xi)| \leq 1 ; \quad(\xi \in D)
$$

satisfying

$$
\begin{equation*}
u(\xi)=\psi(\xi) v(w(\xi)) ; \quad(\xi \in D) \tag{2}
\end{equation*}
$$

Remark 1. (i) We have the conventional definition of subordination if we put $\psi(\xi)=1$ in (2).
(ii) We have the conventional definition of majorization if we put $w(\xi)=\xi$ in (2).

Let $\mathcal{A}$ be the class of all functions of the form

$$
\begin{equation*}
f(\xi)=\xi+\sum_{\mathfrak{K}=2}^{\infty} a_{\mathfrak{K}} \xi^{\mathfrak{K}} ; \quad(\xi \in D), \tag{3}
\end{equation*}
$$

which are analytic in open unit disk $D$, and consider $H_{s}: \mathcal{A} \rightarrow \mathcal{A}$ be an operator such that $\frac{\xi H_{s+1}^{\prime}(f)(\xi)}{H_{s+1}(f)(\xi)}$ is analytic in $D$ with

$$
\left.\frac{\xi H_{s+1}^{\prime}(f)(\xi)}{H_{s+1}(f)(\xi)}\right|_{\xi=0}=\beta+k+\gamma
$$

and satisfies

$$
\begin{equation*}
\xi H_{s+1}^{\prime}(f)(\xi)=k H_{s+1}(f)(\xi)+m H_{s}(f)(\xi), \quad \forall f \in \mathcal{A} \tag{4}
\end{equation*}
$$

for some $\gamma, m, k \in \mathbb{C}$, and $\beta$ is a real number with $\beta>0$ (See 2 ).
Remark 2. (i) If we take $k=-n, m=n+1, \beta=1-\eta$, and $\gamma=\eta+n$ for some integers $n>-1$ and $0 \leq \eta<1$, then the operator $H_{s}$ reduced into the integral operator $\mathcal{I}_{n}$ introduced by Liu and Noor in (16].
(ii) If we take $k=-b, m=1+b, \mu=1-\alpha$ and $\gamma=\alpha+b$, for $b \in \mathbb{C} \backslash Z_{0}^{-}, 0 \leq \alpha<1$, then the operator $H_{s}$ reduced into the Srivastava-Attiya operator $J_{s, b}$, (see [12] and (20]).

Now, using the operator $H_{s}$, we express the following classes of analytic functions.
Definition 1. The function $f \in \mathcal{A}$ is stated to be in the class $S_{k}[E, F ; \mu ; \gamma]$ if and only if

$$
\begin{equation*}
1+\frac{1}{\mu}\left(\frac{\xi\left(H_{s} f(\xi)\right)^{\prime}}{H_{s} f(\xi)}-k-\gamma\right) \prec \frac{1+E \xi}{1+F \xi} \tag{5}
\end{equation*}
$$

with $k, \gamma \in \mathbb{C}, \mu \in \mathbb{C} \backslash\{0\}$ and $-1 \leq F<E \leq 1$.

If we take the value of $k, m, \beta$ and $\gamma$ as defined in Remark (1.2)(i), then this class becomes $S_{n}[E, F ; \mu ; \eta]$ which is defined by Liu and Noor in (16].
Again if we take the value of $k, m, \mu$ and $\gamma$ as defined in Remark (1.2)(ii), then this class becomes $H_{s, b, \alpha}(E, F)$ which is defined by Kutbi and Attiya in 12.
Definition 2. The function $f \in \mathcal{A}$ is stated to be in the class $T_{k}(\theta, \mu, \gamma)$ if and only if

$$
\begin{equation*}
\frac{e^{i \theta}}{\mu+k+\gamma}\left(\frac{\xi\left(H_{s} f(\xi)\right)^{\prime}}{H_{s} f(\xi)}\right) \prec e^{\xi} \cos \theta+i \sin \theta ; \quad(\xi \in D) \tag{6}
\end{equation*}
$$

where $k, \gamma \in \mathbb{C}, \mu \in \mathbb{C} \backslash\{0\}$ and $-\frac{\Pi}{2}<\theta<\frac{\Pi}{2}$.
If we take the value of $k, m, \beta$ and $\gamma$ as defined in Remark (1.2)(i), then this class become as $T_{n}[\theta ; \mu ; \eta]$.
If we take the value of $k, m, \mu$ and $\gamma$ as defined in Remark (1.2)(ii), then this class becomes $T_{b, \alpha}$.

Numerous mathematicians have recently investigated various majorization problems for univalent and multivalent functions as well as meromorphic and multivalent comprising distinct operators and different groups, (see |1, 6, 7, 8, [9, 10, 21, 22).

The majorization problems of the classes $S_{k}[E, F ; \mu ; \gamma]$ and $T_{k}(\theta, \mu, \gamma)$ are explored in this study as follows:

## 2. Main Results

Theorem 1. Assume the function $f \in \mathcal{A}$ and that $g \in S_{k}[E, F ; \mu ; \gamma]$. If $H_{s} f(\xi)$ is majorized by $H_{s} g(\xi)$ in $D$, then

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq\left|H_{s-1} g(\xi)\right|, \quad \text { for } \quad|\xi| \leq \epsilon_{0} \tag{7}
\end{equation*}
$$

where the least positive root of following equation is $\epsilon_{0}$.

$$
\begin{align*}
& |\mu(E-F)+\gamma F| \epsilon^{3}-(2|F|+|\gamma|) \epsilon^{2}-[2+|\mu(E-F)+\gamma F|] \epsilon \\
& \quad+|\gamma|=0 \tag{8}
\end{align*}
$$

and $-1 \leq F<E \leq 1, k, \gamma, m \in \mathbb{C}, \mu \in \mathbb{C} \backslash\{0\}$.
Proof. Since $g \in S_{k}[E, F ; \mu ; \gamma]$ then, from (5) and definition of majorization

$$
1+\frac{1}{\mu}\left(\frac{\xi\left(H_{s}^{\prime} g(\xi)\right)}{H_{s} g(\xi)}-k-\gamma\right)=\frac{1+E w(\xi)}{1+F w(\xi)}
$$

with $w(0)=0$ and $|w(\xi)| \leq|\xi|<1, \quad \forall \xi \in D$.
Now, from the above equality

$$
\begin{equation*}
\frac{\xi\left(H_{s}^{\prime} g(\xi)\right)}{H_{s} g(\xi)}=\frac{(k+\gamma)+(\mu(E-F)+(k+\gamma) F) w(\xi)}{1+F w(\xi)} \tag{9}
\end{equation*}
$$

Using the relation (4), that is,

$$
\xi\left(H_{s}^{\prime} g(\xi)\right)=k H_{S} g(\xi)+m H_{S-1} g(\xi)
$$

for $k, m \in \mathbb{C}$, we have from (9) as

$$
\frac{H_{S-1} g(\xi)}{H_{s} g(\xi)}=\frac{\gamma+(\mu(E-F)+\gamma F) w(\xi)}{m(1+F w(\xi))}
$$

which implies that

$$
\begin{equation*}
\left|H_{s} g(\xi)\right| \leq \frac{|m|(1+|F||\xi|)\left|H_{S-1} g(\xi)\right|}{|\gamma|-|\mu(E-F)+\gamma F||\xi|} . \tag{10}
\end{equation*}
$$

As $H_{s} f(\xi)$ is majorized by $H_{S} g(\xi)$ in open unit disk $D$, then

$$
\begin{equation*}
H_{s} f(\xi)=\psi(\xi) H_{s} g(\xi) \tag{11}
\end{equation*}
$$

Multiplying (11) by $\xi$ after differentiating with respect to $\xi$, we get

$$
\xi\left(H_{s}^{\prime} f(\xi)\right)=\xi \psi(\xi)\left(H_{s}^{\prime} g(\xi)\right)+\xi \psi^{\prime}(\xi) H_{s} g(\xi)
$$

on using relation (4), we have

$$
m H_{s-1} f(\xi)=\xi \psi^{\prime}(\xi) H_{s} g(\xi)+m \psi(\xi) H_{s-1} g(\xi)
$$

that implies

$$
\begin{equation*}
|m|\left|H_{s-1} f(\xi)\right| \leq|\xi|\left|\psi^{\prime}(\xi)\right|\left|H_{s} g(\xi)\right|+|m||\psi(\xi)|\left|H_{s-1} g(\xi)\right| \tag{12}
\end{equation*}
$$

As a consequence, considering that the $\psi$ (Schwarz function) meets the inequality, (see 18)

$$
\begin{equation*}
\left|\psi^{\prime}(\xi)\right| \leq \frac{1-|\psi(\xi)|^{2}}{1-|\xi|^{2}} ; \quad(\xi \in D) \tag{13}
\end{equation*}
$$

on using (10) and (13) in (12), we have

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq\left[\frac{|\xi|\left(1-|\psi(\xi)|^{2}\right)(1+|F||\xi|)}{\left(1-|\xi|^{2}\right)(|\gamma|-|\mu(E-F)+\gamma F||\xi|)}+|\psi(\xi)|\right]\left|H_{s-1} g(\xi)\right| \tag{14}
\end{equation*}
$$

Setting $|\xi|=\epsilon,|\psi(\xi)|=\kappa$, then inequality (14) leads to

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq \frac{\zeta(\epsilon, \kappa)\left|H_{s-1} g(\xi)\right|}{\left(1-\epsilon^{2}\right)(|\gamma|-|\mu(E-F)+\gamma F| \epsilon)} \tag{15}
\end{equation*}
$$

where

$$
\zeta(\epsilon, \kappa)=\epsilon\left(1-\kappa^{2}\right)(1+|F| \epsilon)+\kappa\left(1-\epsilon^{2}\right)[|\gamma|-|\mu(E-F)+\gamma F| \epsilon] .
$$

Then, from (15)

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq \mathfrak{T}(\epsilon, \kappa)\left|H_{s-1} g(\xi)\right| \tag{16}
\end{equation*}
$$

where

$$
\begin{equation*}
\mathfrak{T}(\epsilon, \kappa)=\frac{\zeta(\epsilon, \kappa)}{\left(1-\epsilon^{2}\right)(|\gamma|-|\mu(E-F)+\gamma F| \epsilon)} \tag{17}
\end{equation*}
$$

from relation (16), in an attempt to prove our result, we have to specify

$$
\begin{array}{rlrl}
\epsilon_{0} & =\max \{\epsilon \in[0,1) ; \mathfrak{T}(\epsilon, \kappa) \leq 1 ; & & \forall \kappa \in[0,1]\} \\
& =\max \{\epsilon \in[0,1) ; G(\epsilon, \kappa) \geq 0 ; & \forall \kappa \in[0,1]\}
\end{array}
$$

where

$$
\begin{aligned}
G(\epsilon, \kappa)= & \left(1-\epsilon^{2}\right)(1-\kappa)[|\gamma|-|\mu(E-F)+\gamma F| \epsilon] \\
& -\epsilon\left(1-\kappa^{2}\right)(1+|F| \epsilon)
\end{aligned}
$$

A simple calculation shows that the $G(\epsilon, \kappa) \geq 0$ inequality is equivalent to

$$
\begin{aligned}
u(\epsilon, \kappa)= & {[|\gamma|-|\mu(E-F)+\gamma F| \epsilon]\left(1-\epsilon^{2}\right) } \\
& -\epsilon(1+\kappa)(1+|F| \epsilon) \geq 0
\end{aligned}
$$

while the function $u(\epsilon, \kappa)$ has a least value at $\kappa=1$, i.e.
$\min \{u(\epsilon, \kappa): \kappa \in[0,1]\}=u(\epsilon, 1)=v(\epsilon)$,
where

$$
\begin{aligned}
v(\epsilon)= & |\mu(E-F)+\gamma F| \epsilon^{3}-(2|F|+|\gamma|) \epsilon^{2} \\
& -[2+|\mu(E-F)+\gamma F|] \epsilon+|\gamma|=0
\end{aligned}
$$

it follows that $v(\epsilon) \geq 0 ; \quad \forall \epsilon \in\left[0, \epsilon_{0}\right]$, where $\epsilon_{0}=\epsilon_{0}(\mu, \gamma, E, F)$ is the least positive root of equation (8), which proves the conclusion of (7).

Theorem 2. Assume the function $f \in \mathcal{A}$ and that $g \in T_{k}(\theta, \mu, \gamma)$. If $H_{s} f(\xi)$ is majorized by $H_{s} g(\xi)$ in $D$, therefore

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq\left|H_{s-1} g(\xi)\right| \quad \text { for } \quad|\xi| \leq \epsilon_{1} \tag{18}
\end{equation*}
$$

where the least positive root of following equation is $\epsilon_{1}$.
$\epsilon^{2}\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)+2 \epsilon|\sec \theta|-\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)=0$,
and $\gamma, k \in \mathbb{C},-\frac{\Pi}{2}<\theta<\frac{\Pi}{2}, \mu \in \mathbb{C} \backslash\{0\}$.
Proof. Since, $g \in T_{k}(\theta, \mu, \gamma)$ then, from (1) and the subordination relation

$$
\begin{equation*}
\frac{e^{i \theta}}{\mu+k+\gamma}\left(\frac{\xi\left(H_{s}^{\prime} g(\xi)\right)}{H_{s} g(\xi)}\right)=e^{w(\xi)} \cos \theta+i \sin \theta \tag{20}
\end{equation*}
$$

with $w(0)=0$ and $|w(\xi)| \leq 1 \quad \forall \xi \in D$.
From (20), we have

$$
\begin{equation*}
\frac{\xi H_{s}^{\prime} g(\xi)}{H_{s} g(\xi)}=(\mu+k+\gamma)\left(\frac{e^{w(\xi)}+i \tan \theta}{1+i \tan \theta}\right) \tag{21}
\end{equation*}
$$

Now, using (4) in (21), for $\gamma, m, k \in \mathbb{C}$ and $\mu \in \mathbb{C} \backslash\{0\}$, we have the following.

$$
\frac{H_{s-1} g(\xi)}{H_{s} g(\xi)}=\frac{(\mu+k+\gamma) e^{w(\xi)}-k+(\gamma+\mu) i \tan \theta}{m(1+i \tan \theta)}
$$

which implies that

$$
\begin{equation*}
\left|H_{s} g(\xi)\right| \leq \frac{|m||\sec \theta|}{\left(|\mu+k+\gamma| e^{|\xi|}-|k|-|\mu+\gamma||\tan \theta|\right)}\left|H_{s-1} g(\xi)\right| \tag{22}
\end{equation*}
$$

Now, since $H_{s} f(\xi)$ is majorized by $H_{s} g(\xi)$ in $D$, we have

$$
\begin{equation*}
H_{s} f(\xi)=\psi(\xi) H_{s} g(\xi) \tag{23}
\end{equation*}
$$

Multiplying (23) by $\xi$ after differentiating with respect to $\xi$, we get

$$
\xi\left(H_{s}^{\prime} f(\xi)\right)=\xi \psi(\xi)\left(H_{s}^{\prime} g(\xi)\right)+\xi \psi^{\prime}(\xi) H_{s} g(\xi)
$$

on using relation (4), we have

$$
m H_{s-1} f(\xi)=\xi \psi^{\prime}(\xi) H_{s} g(\xi)+m \psi(\xi) H_{s-1} g(\xi)
$$

that implies

$$
\begin{equation*}
|m|\left|H_{s-1} f(\xi)\right| \leq|\xi|\left|\psi^{\prime}(\xi)\right|\left|H_{s} g(\xi)\right|+|m||\psi(\xi)|\left|H_{s-1} g(\xi)\right| \tag{24}
\end{equation*}
$$

As a consequence, considering that the $\psi$ (Schwarz function) meets the inequality, (see 18)

$$
\begin{equation*}
\left|\psi^{\prime}(\xi)\right| \leq \frac{1-|\psi(\xi)|^{2}}{1-|\xi|^{2}} ; \quad(\xi \in D) \tag{25}
\end{equation*}
$$

using (22) and (25) in (24), we have

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq\left(\frac{|\xi|\left(1-|\psi(\xi)|^{2}\right)|\sec \theta|}{\left(1-|\xi|^{2}\right)\left(|\mu+k+\gamma| e^{|\xi|}-|k|-|\mu+\gamma||\tan \theta|\right)}+|\psi(\xi)|\right)\left|H_{s-1} g(\xi)\right| \tag{26}
\end{equation*}
$$

Setting $|\xi|=\epsilon,|\psi(\xi)|=\kappa \quad(0 \leq \kappa \leq 1)$, then inequality 26) leads to

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq \frac{\zeta_{1}(\epsilon, \kappa)}{\left(1-\epsilon^{2}\right)\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)}\left|H_{s-1} g(\xi)\right| \tag{27}
\end{equation*}
$$

where

$$
\zeta_{1}(\epsilon, \kappa)=\epsilon\left(1-\kappa^{2}\right)|\sec \theta|+\kappa\left(1-\epsilon^{2}\right)\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)
$$

Then, from (27)

$$
\begin{equation*}
\left|H_{s-1} f(\xi)\right| \leq \mathfrak{T}_{1}(\epsilon, \kappa)\left|H_{s-1} g(\xi)\right| \tag{28}
\end{equation*}
$$

where

$$
\begin{equation*}
\mathfrak{T}_{1}(\epsilon, \kappa)=\frac{\zeta_{1}(\epsilon, \kappa)}{\left(1-\epsilon^{2}\right)\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)} \tag{29}
\end{equation*}
$$

From relation (28), in order to prove our result, we have to specify

$$
\epsilon_{1}=\max \left\{\epsilon \in[0,1) ; \mathfrak{T}_{1}(\epsilon, \kappa) \leq 1 \quad \forall \kappa \in[0,1]\right\}
$$

$$
=\max \left\{\epsilon \in[0,1) ; G_{1}(\epsilon, \kappa) \geq 0 \quad \forall \kappa \in[0,1]\right\}
$$

where

$$
G_{1}(\epsilon, \kappa)=\left(1-\epsilon^{2}\right)(1-\kappa)\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)-\epsilon\left(1-\kappa^{2}\right)|\sec \theta| .
$$

A quick calculation illustrates that the inequality $G_{1}(\epsilon, \kappa) \geq 0$ is equivalent to

$$
u_{1}(\epsilon, \kappa)=\left(1-\epsilon^{2}\right)\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)-\epsilon(1+\kappa)|\sec \theta| \geq 0
$$

while the function $u_{1}(\epsilon, \kappa)$ takes its lowest value at $\kappa=1$, that is,

$$
\min \left\{u_{1}(\epsilon, \kappa): \kappa \in[0,1]\right\}=u_{1}(\epsilon, 1)=v_{1}(\epsilon)
$$

where

$$
v_{1}(\epsilon)=\left(1-\epsilon^{2}\right)\left(|\mu+k+\gamma| e^{\epsilon}-|k|-|\mu+\gamma||\tan \theta|\right)-2 \epsilon|\sec \theta|=0
$$

It follows that $v_{2}(\epsilon) \geq 0 \quad \forall \epsilon \in\left[0, \epsilon_{1}\right]$, where $\epsilon_{1}=\epsilon_{1}(\theta, \gamma, \mu, k)$ is the least positive root of equation (19), which proves the conclusion of (18).

## 3. Corollaries and Consequences

Corollary 1. Assume the function $f \in \mathcal{A}$ and that $g \in S_{n}[E, F ; \mu ; \eta]$. If $\mathcal{I}_{n} f(\xi)$ is majorized by $\mathcal{I}_{n} g(\xi)$ in $D$, then

$$
\begin{equation*}
\left|\mathcal{I}_{n-1} f(\xi)\right| \leq\left|\mathcal{I}_{n-1} g(\xi)\right| \quad \text { for } \quad|\xi| \leq \epsilon_{2} \tag{30}
\end{equation*}
$$

where the least positive root of following equation is $\epsilon_{2}$.

$$
\begin{align*}
& |\mu E+(n+\eta-\mu) F| \epsilon^{3}-(2|F|+|n+\eta|) \epsilon^{2}-(2+|\mu E+(\eta+n-\mu) F|) \epsilon+|\eta+n|=0,  \tag{31}\\
& \text { and }-1 \leq F<E \leq 1, \mu \in \mathbb{C} \backslash\{0\}, n>-1,0 \leq \eta<1,
\end{align*}
$$

Corollary 2. Assume the function $f \in \mathcal{A}$ and that $g \in T_{n}[\theta ; \mu ; \eta]$. If $\mathcal{I}_{n} f(\xi)$ is majorized by $\mathcal{I}_{n} g(\xi)$ in $D$, then

$$
\begin{equation*}
\left|\mathcal{I}_{n-1} f(\xi)\right| \leq\left|\mathcal{I}_{n-1} g(\xi)\right| \quad \text { for } \quad|\xi| \leq \epsilon_{3}, \tag{32}
\end{equation*}
$$

where the least positive root of following equation is $\epsilon_{3}$.
$\left(|\mu+\eta| e^{\epsilon}-|n|-|\mu+\eta+n||\tan \theta|\right) \epsilon^{2}-2|\sec \theta| \epsilon-\left(-|n|-|\mu+\eta+n||\tan \theta|+|\mu+\eta| e^{\epsilon}\right)=0$,
and $n>-1,0 \leq \eta<1,-\frac{\pi}{2}<\theta<\frac{\pi}{2}$..
Corollary 3. Assume the function $f \in \mathcal{A}$ and that $g \in H_{s, b, \alpha}(E, F)$. If $J_{s, b} f(\xi)$ is majorized by $J_{s, b} g(\xi)$ in $D$, then

$$
\begin{equation*}
\left|J_{s-1, b} f(\xi)\right| \leq\left|J_{s-1, b} g(\xi)\right| \quad \text { for } \quad|\xi| \leq \epsilon_{4} \tag{34}
\end{equation*}
$$

where the least positive root of following equation is $\epsilon_{4}$.
$|(1-\alpha) E+(2 \alpha+b-1)| \epsilon^{3}-(2|F|+|\alpha+b|) \epsilon^{2}-(2+|(1-\alpha) E+(2 \alpha+b-1) F|) \epsilon$ $+|\alpha+b|=0$,
and $-1 \leq F<E \leq 1, b \in \mathbb{C} \backslash Z_{0}^{-}, 0 \leq \alpha<1$.

Corollary 4. Assume the function $f \in \mathcal{A}$ and that $g \in T_{b, \alpha}$. If $J_{s, b} f(\xi)$ is majorized by $J_{s, b} g(\xi)$ in $D$, then

$$
\begin{equation*}
\left|J_{s-1, b} f(\xi)\right| \leq\left|J_{s-1, b} g(\xi)\right| \quad \text { for } \quad|\xi| \leq \epsilon_{5}, \tag{36}
\end{equation*}
$$

where the least positive root of following equation is $\epsilon_{5}$.

$$
\begin{equation*}
\left(e^{\epsilon}-|1+b||\tan \theta|-|b|\right) \epsilon^{2}+2|\sec \theta| \epsilon-\left(e^{\epsilon}-|b|-|1+b||\tan \theta|\right)=0, \tag{37}
\end{equation*}
$$

and $b \in \mathbb{C} \backslash Z_{0}^{-}, 0 \leq \alpha<1,-\frac{\pi}{2}<\theta<\frac{\pi}{2}$.

Author Contribution Statements Each author contributes equally to the preparation of the manuscript.

Declaration of Competing Interests The author declares that he has no competing interests.

Acknowledgements The authors thank the reviewer and the editor for their constructive comments to improve the article.

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$$
\begin{aligned}
x_{21} & =c_{1} a_{21} \\
x_{22} & =c_{0}+c_{1} a_{22}
\end{aligned}
$$

In equation (6), let us write $\lambda_{1}$ and $\lambda_{2}$ instead of $A$ and find $c_{0}$ and $c_{1}$ values

$$
\begin{aligned}
& c_{0}=\frac{(\beta+2)^{n} f\left(x(\alpha+2)^{n}\right)-(\alpha+2)^{n} f\left(x(\beta+2)^{n}\right)}{(\beta+2)^{n}-(\alpha+2)^{n}} \\
& c_{1}=\frac{f\left(x(\beta+2)^{n}\right)-f\left(x(\alpha+2)^{n}\right)}{(\beta+2)^{n}-(\alpha+2)^{n}}
\end{aligned}
$$

and then

$$
\begin{aligned}
& x_{11}=\left[\left(a_{11}-\lambda_{1}\right) f\left(\lambda_{2}\right)-\left(a_{11}-\lambda_{2}\right) f\left(\lambda_{1}\right)\right] /\left(\lambda_{2}-\lambda_{1}\right), \\
& x_{12}=a_{12}\left[f\left(\lambda_{2}\right)-f\left(\lambda_{1}\right)\right] /\left(\lambda_{2}-\lambda_{1}\right) \\
& x_{21}=a_{21}\left[f\left(\lambda_{2}\right)-f\left(\lambda_{1}\right)\right] /\left(\lambda_{2}-\lambda_{1}\right) \\
& x_{22}=\left[\left(a_{22}-\lambda_{1}\right) f\left(\lambda_{2}\right)-\left(a_{22}-\lambda_{2}\right) f\left(\lambda_{1}\right)\right] /\left(\lambda_{2}-\lambda_{1}\right) .
\end{aligned}
$$

Lemma 2. Let $k$ and $n$ be arbitrary positive integers. For $x$ an arbitrary quantity, let us consider the matrix $x M_{k}^{n}$ having eigenvalues

$$
\begin{aligned}
& \lambda_{1}=x(\alpha+2)^{n} \\
& \lambda_{2}=x(\beta+2)^{n}
\end{aligned}
$$

Proof. It is easily seen by induction.

To express the $y_{i j}$ components of $Y=\left[y_{i j}\right]=f\left(x M_{k}^{n}\right)$ in separate formulas, we can give the following theorem with

$$
\lambda:=\frac{(\beta+2)^{n} f\left(x(\alpha+2)^{n}\right)-(\alpha+2)^{n} f\left(x(\beta+2)^{n}\right)}{(\beta+2)^{n}-(\alpha+2)^{n}}
$$

and

$$
\phi:=\frac{f\left(x(\beta+2)^{n}\right)-f\left(x(\alpha+2)^{n}\right)}{(\beta+2)^{n}-(\alpha+2)^{n}}
$$

Theorem 3. Let $k$ and $n$ be arbitrary positive integers.
i) If $n$ is even and $k$ is odd, then

$$
Y=\frac{5^{\frac{n}{2}}}{F_{k}}\left[\begin{array}{cc}
\lambda F_{k}+\phi F_{n+k} & \phi F_{n} \\
\phi F_{n} & \lambda F_{k}+\phi F_{n-k}
\end{array}\right]
$$

ii) If $n$ is odd and $k$ is odd, then

$$
Y=\frac{5^{\frac{n-1}{2}}}{F_{k}}\left[\begin{array}{cc}
\lambda F_{k}+\phi L_{n+k} & \phi L_{n} \\
\phi L_{n} & \lambda F_{k}+\phi L_{n-k}
\end{array}\right] .
$$

iii) If $n$ is odd and $k$ is even, then

$$
Y=\frac{5^{\frac{n-1}{2}}}{L_{k}}\left[\begin{array}{cc}
\lambda L_{k}+5 \phi F_{n+k} & \sqrt{5} \phi L_{n} \\
\sqrt{5} \phi L_{n} & \lambda L_{k}+5 \phi F_{n-k}
\end{array}\right]
$$

iv) If $n$ is even and $k$ is even, then

$$
Y=\frac{5^{\frac{n}{2}}}{L_{k}}\left[\begin{array}{cc}
\lambda L_{k}+\phi L_{n+k} & \sqrt{5} \phi F_{n} \\
\sqrt{5} \phi F_{n} & \lambda L_{k}+\phi L_{n-k}
\end{array}\right]
$$

Proof. Taking $x M_{k}^{n}$ as matrix $A$ in equation (6) and applying the above steps using Lemma 2 the desired result is obtained.

Theorem 4. If $f$ is the matrix inversion function then

$$
\left(x M_{k}^{n}\right)^{-1}= \begin{cases}\frac{5^{\frac{-n-1}{2}}}{x F_{k}}\left[\begin{array}{cc}
L_{n-k} & -L_{n} \\
-L_{n} & L_{n+k}
\end{array}\right], & \text { if } k \text { is odd and } n \text { is odd, } \\
\frac{5^{\frac{-n}{2}}}{x F_{k}}\left[\begin{array}{cc}
F_{n-k} & -F_{n} \\
-F_{n} & F_{n+k}
\end{array}\right], & \text { if } k \text { is odd and } n \text { is even, } \\
\frac{5^{\frac{-n}{2}}}{x L_{k}}\left[\begin{array}{cc}
L_{n-k} & -F_{n} \sqrt{5} \\
-F_{n} \sqrt{5} & L_{n+k}
\end{array}\right], & \text { if } k \text { is even and } n \text { is even, } \\
\frac{5^{\frac{-n-1}{2}}}{x L_{k}}\left[\begin{array}{cc}
5 F_{n-k} & -L_{n} \sqrt{5} \\
-L_{n} \sqrt{5} & 5 F_{n+k}
\end{array}\right], & \text { if } k \text { is even and } n \text { is odd. }\end{cases}
$$

Proof. It can be easily seen using the identity $\left(x M_{k}^{n}\right)^{-1}=\frac{1}{x} M_{k}^{-n},(x \neq 0)$.

## 3. Relations with Some Finite Series

In this section, sums of some finite series containing $F_{n}$ and $L_{n}$ are found using some properties of the Lucas-type Cholesky algorithm matrix $M_{k}$.

Lemma 3. If $k$ is a positive integer, then

$$
\begin{equation*}
M_{k}^{2}=5 M_{k}-5 I \tag{7}
\end{equation*}
$$

and

$$
\begin{equation*}
M_{k}^{-1}=I-\frac{1}{5} M_{k} \tag{8}
\end{equation*}
$$

Proof. Using equation (1), it easily be obtained from equations Theorem 1 and Theorem 2.

Lemma 4. If $x$ is an arbitrary quantity with the constraints $x \neq \frac{1}{\alpha^{n}}$ and $x \neq \frac{1}{\beta^{n}}$ then

$$
\left(x M_{k}^{n}-I\right)^{-1}= \begin{cases}\frac{\left(5^{\frac{n+1}{2}} F_{n} x-1\right) I-x M_{k}^{n}}{5^{n} x^{2}-5^{\frac{n+1}{2}} F_{n} x+1}, & \text { if } n \text { is odd } \\ \frac{\left(5^{\frac{n}{2}} L_{n} x-1\right) I-x M_{k}^{n}}{5^{n} x^{2}-5^{\frac{n}{2}} L_{n} x+1}, & \text { if } n \text { is even }\end{cases}
$$

Proof. It can be easily seen using equations (2), (3), (4) and Lemma 3 and the following equations

$$
\begin{aligned}
L_{k+n}-5 F_{n} F_{k} & =-L_{n-k} \\
F_{k+n}-F_{k} L_{n} & =-F_{n-k}
\end{aligned} \quad \text { if } k \text { is odd and } n \text { is odd } \text { odd }[15, p .111,83 .], ~ \text { is even [15, p. 118, 58.], }, ~ \begin{gathered}
\text { if } k \text { is even and } n \text { is even [15, p. 111, 83.], } \\
L_{k+n}-L_{n} L_{k}=-L_{n-k} \\
F_{k+n}-L_{k} F_{n}=-F_{n-k} \quad \text { if } k \text { is even and } n \text { is odd [15, p. 118, 58.]. }
\end{gathered}
$$

Lemma 5. For positive numbers $k$ and $n$ the following equality holds

$$
M_{k}^{n}=\sum_{j=0}^{n} 5^{-j}\binom{n}{j} M_{k}^{2 j}
$$

Proof. From equation (7) we can write $\left(M_{k}^{2}+5 I\right)^{n}=\left(5 M_{k}\right)^{n}$, from which the proof can be obtained by using the binomial expansion.

Theorem 5. i) Let $n$ be a nonnegative even integer and $k$ be an arbitrary positive integer. Then we have

$$
\begin{aligned}
& F_{n \mp k}=5^{-\frac{n}{2}} \sum_{j=0}^{n}\binom{n}{j} F_{2 j \mp k}, \\
& L_{n \mp k}=5^{-\frac{n}{2}} \sum_{j=0}^{n}\binom{n}{j} L_{2 j \mp k} .
\end{aligned}
$$

ii) Let $n$ be a nonnegative odd integer and $k$ be an arbitrary positive integer. Then we have

$$
\begin{aligned}
& F_{n \mp k}=5^{\frac{-n-1}{2}} \sum_{j=0}^{n}\binom{n}{j} F_{2 j \mp k} \\
& L_{n \mp k}=5^{\frac{-n+1}{2}} \sum_{j=0}^{n}\binom{n}{j} L_{2 j \mp k} .
\end{aligned}
$$

Proof. If $n$ is even positive integer and $k$ is odd positive integer, then from Theorem 2 and Lemma 5,

$$
M_{k}^{n}=\frac{5^{\frac{n}{2}}}{F_{k}}\left[\begin{array}{cc}
F_{n+k} & F_{n} \\
F_{n} & F_{n-k}
\end{array}\right]=\sum_{j=0}^{n} 5^{-j}\binom{n}{j} \frac{5^{j}}{F_{k}}\left[\begin{array}{cc}
F_{2 j+k} & F_{2 j} \\
F_{2 j} & F_{2 j-k}
\end{array}\right]
$$

hence,

$$
5^{\frac{n}{2}}\left[\begin{array}{cc}
F_{n+k} & F_{n} \\
F_{n} & F_{n-k}
\end{array}\right]=\left[\begin{array}{cc}
\sum_{j=0}^{n}\binom{n}{j} F_{2 j+k} & \sum_{j=0}^{n}\binom{n}{j} F_{2 j} \\
\sum_{j=0}^{n}\binom{n}{j} F_{2 j} & \sum_{j=0}^{n}\binom{n}{j} F_{2 j-k}
\end{array}\right],
$$

therefore,

$$
F_{n \mp k}=5^{-\frac{n}{2}} \sum_{j=0}^{n}\binom{n}{j} F_{2 j \mp k} .
$$

Other equations are obtained in a similar way.
Lemma 6. For positive integers $k, n, s$ the following equality holds

$$
M_{k}^{2 n+s}=5^{n} \sum_{j=0}^{n}(-1)^{n+j}\binom{n}{j} M_{k}^{s+j}
$$

Proof. From equation (7), we can write

$$
\begin{equation*}
\left(5 M_{k}-5 I\right)^{n} M_{k}^{s}=M_{k}^{2 n+s} \tag{9}
\end{equation*}
$$

from which the proof can be obtained by using the binomial expansion.

Theorem 6. For positive integers $n$ and $s$ the following equality holds

$$
\begin{aligned}
& L_{2 n+s}=\sum_{j=0}^{n}\binom{n}{j} \begin{cases}(-1)^{n+1} 5^{\frac{j+1}{2}} F_{s+j}, & \text { if } j \text { is odd }, \\
(-1)^{n} 5^{\frac{j}{2}} L_{s+j}, & \text { if } j \text { is even },\end{cases} \\
& F_{2 n+s}=\sum_{j=0}^{n}\binom{n}{j} \begin{cases}(-1)^{n+1} 5^{\frac{j-1}{2}} L_{s+j}, & \text { if } j \text { is odd, } \\
(-1)^{n} 5^{\frac{j}{2}} F_{s+j}, & \text { if } j \text { is even. } .\end{cases}
\end{aligned}
$$

Proof. It can be easily seen with Lemma 6 and Theorem 2.

Theorem 7. For positive integers $k$ and $n$ the followings holds

$$
\begin{aligned}
& L_{n \pm k}=\left\{\begin{array}{l}
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
-5^{\frac{n-j}{2}} L_{j \mp k}, & \text { if } j \text { is odd, } \\
5^{\frac{n-j+1}{2}} F_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is odd and } n\right. \text { is odd, } \\
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
5^{\frac{n-j}{2}} L_{j \mp k}, & \text { if } j \text { is odd, } \\
-5^{\frac{n-j+1}{2}} F_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is even and } n\right. \text { is odd, } \\
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
5^{\frac{n-j+1}{2}} F_{j \mp k}, & \text { if } j \text { is odd, } \\
-5^{\frac{n-j}{2}} L_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is odd and } n\right. \text { is even, }
\end{array} \quad \begin{array}{l}
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
-5^{\frac{n-j+1}{2}} F_{j \mp k}, & \text { if } j \text { is odd, } \\
5^{\frac{n-j}{2}} L_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is even and } n\right. \text { is even, }
\end{array}\right. \\
& F_{n \pm k}=\left\{\begin{array}{l}
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
5^{\frac{n-j}{2}} F_{j \mp k}, & \text { if } j \text { is odd, } \\
-5^{\frac{n-j-1}{2}} L_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is odd and } n \text { is odd, },\right. \\
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
-5^{\frac{n-j}{2}} F_{j \mp k}, & \text { if } j \text { is odd, } \\
5^{\frac{n-j-1}{2}} L_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is even and } n\right. \text { is odd, } \\
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
-5^{\frac{n-j-1}{2}} L_{j \mp k}, & \text { if } j \text { is odd, } \\
5^{\frac{n-j}{2}} F_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is odd and } n\right. \text { is even, }
\end{array} \quad \begin{array}{l}
\sum_{j=0}^{n}\binom{n}{j}\left\{\begin{array}{ll}
5^{\frac{n-j-1}{2}} L_{j \mp k}, & \text { if } j \text { is odd, } \\
-5^{\frac{n-j}{2}} F_{j \mp k}, & \text { if } j \text { is even, }
\end{array} \quad \text { if } k \text { is even and } n\right. \text { is even. }
\end{array}\right.
\end{aligned}
$$

Proof. Using equation (8) we can write $\left(I-\frac{1}{5} M_{k}\right)^{n}=\left(M_{k}^{n}\right)^{-1}$. Here,

$$
\left(I-\frac{1}{5} M_{k}\right)^{n}=\sum_{j=0}^{n}\binom{n}{j}(-1)^{j} \frac{1}{5^{j}} M_{k}^{j}=\left(M_{k}^{n}\right)^{-1}
$$

Let $n, k$ be odd positive integers.

$$
\begin{gathered}
\sum_{\substack{j=0 \\
j \text { odd }}}^{n}\binom{n}{j}(-1)^{j} \frac{1}{5^{j}} \frac{5^{\frac{j-1}{2}}}{F_{k}}\left[\begin{array}{cc}
L_{j+k} & L_{j} \\
L_{j} & L_{j-k}
\end{array}\right]+\sum_{\substack{j=0 \\
j \text { even }}}^{n}\binom{n}{j}(-1)^{j} \frac{1}{5^{j}} \frac{5^{\frac{j}{2}}}{F_{k}}\left[\begin{array}{cc}
F_{j+k} & F_{j} \\
F_{j} & F_{j-k}
\end{array}\right] \\
=\frac{5^{\frac{-(n+1)}{2}}}{F_{k}}\left[\begin{array}{cc}
L_{n-k} & -L_{n} \\
-L_{n} & L_{n+k}
\end{array}\right]
\end{gathered}
$$

hence,

$$
\left[\begin{array}{cc}
L_{n-k} & -L_{n} \\
-L_{n} & L_{n+k}
\end{array}\right]=\sum_{j=0}^{n}\binom{n}{j}(-1)^{j}\left\{\begin{array}{cl}
5^{\frac{n-j}{2}}\left[\begin{array}{cc}
L_{j+k} & L_{j} \\
L_{j} & L_{j-k}
\end{array}\right], & \text { if } j \text { is odd }, \\
5^{\frac{n-j+1}{2}}\left[\begin{array}{cc}
F_{j+k} & F_{j} \\
F_{j} & F_{j-k}
\end{array}\right], & \text { if } j \text { is even }
\end{array}\right.
$$

from which the following result is obtained

$$
\begin{aligned}
& L_{n-k}=\sum_{j=0}^{n}\binom{n}{j} \begin{cases}-5^{\frac{n-j}{2}} L_{j+k}, & \text { if } j \text { is odd }, \\
5^{\frac{n-j+1}{2}} F_{j+k}, & \text { if } j \text { is even, }\end{cases} \\
& L_{n+k}=\sum_{j=0}^{n}\binom{n}{j} \begin{cases}-5^{\frac{n-j}{2}} L_{j-k}, & \text { if } j \text { is odd, } \\
5^{\frac{n-j+1}{2}} F_{j-k}, & \text { if } j \text { is even. }\end{cases}
\end{aligned}
$$

Other equations are obtained in a similar way.
Theorem 8. Let $h, k$ and $n$ be positive integers and

$$
\theta(n):=5^{\frac{n+1}{2}} F_{n} x-1, \quad \vartheta(n):=5^{\frac{n}{2}} L_{n} x-1
$$

i) If $n$ is odd, then

$$
\begin{aligned}
\sum_{j=0}^{h} x^{j} M_{k}^{n j} & =\frac{\theta(n) I-x M_{k}^{n}}{5^{n} x^{2}-\theta(n)}\left(x^{h+1} M_{k}^{n(h+1)}-I\right) \\
& =-\frac{x^{h+2} M_{k}^{n(h+2)}-x M_{k}^{n}-\theta(n)\left(x M_{k}^{n}\right)^{h+1}+\theta(n) I}{5^{n} x^{2}-\theta(n)}
\end{aligned}
$$

ii) If $n$ is even, then

$$
\begin{aligned}
\sum_{j=0}^{h} x^{j} M_{k}^{n j} & =\frac{\vartheta(n) I-x M_{k}^{n}}{5^{n} x^{2}-\vartheta(n)}\left(x^{h+1} M_{k}^{n(h+1)}-I\right) \\
& =-\frac{x^{h+2} M_{k}^{n(h+2)}-x M_{k}^{n}-\vartheta(n)\left(x M_{k}^{n}\right)^{h+1}+\vartheta(n) I}{5^{n} x^{2}-\vartheta(n)}
\end{aligned}
$$

Proof.

$$
\begin{equation*}
\left(x A^{n}-I\right) \sum_{j=0}^{h} x^{j} A^{n j}=x^{h+1} A^{n(h+1)}-I \tag{10}
\end{equation*}
$$

is valid for every square matrix A. Using equation (10) and Lemma 4, i) and ii) can easily be shown.

Theorem 9. Let $n$ and $s$ be arbitrary integers where $x \neq \frac{1}{\alpha^{n}}$ and $x \neq \frac{1}{\beta^{n}}$, the following equations are satisfied:
i)

$$
\sum_{j=0}^{h} x^{j} F_{n j+s}=\frac{(-1)^{n-1} x^{h+2} F_{n h+s}+x^{h+1} F_{n(h+1)+s}-(-1)^{s} x F_{n-s}-F_{s}}{(-1)^{n-1} x^{2}+L_{n} x-1}
$$


$\qquad$
$\qquad$

$\qquad$




Theorem 1. Let be the free metabelian Leibniz algebra with a generating set $n$. Then is generated by the general linear group together with the inner automorphisms and the following -automorphisms

```
where
        and
        n,
        j j j
where is generated by the elements of the form
                where
                                    n ,
where
depends on t ,
where
Theorem 2. Let
Then
Proof.



Theorem 3.
Proof. \(\square\)
\(i\)
Case 1. \({ }_{i}{ }_{i}\)
Case 2.

Case 3.
\(n\)

C ase 4.

C ase 5.
\(j \quad k\)
\(j\)

Declaration of Competing Interests
Acknowledgements
http://communications.science.ankara.edu.tr

\title{
AFFINE MAPPINGS AND MULTIPLIERS FOR WEIGHTED ORLICZ SPACES OVER THE AFFINE GROUP \(\mathbb{R}_{+} \times \mathbb{R}\)
}

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}

\begin{abstract}
Let \(\mathbb{A}=\mathbb{R}_{+} \times \mathbb{R}\) be the affine group with a right Haar measure \(\mu, \omega\) be a weight function on \(\mathbb{A}\) and \(\Phi\) be a Young function. We characterize the affine continuous mappings on the subsets of \(L^{\Phi}(\mathbb{A}, \omega)\). Moreover we show that there exists an isometric isomorphism between the multiplier of the pair \(\left(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}), L^{1}(\mathbb{A})\right)\) and the space of bounded measures \(M(\mathbb{A})\).
\end{abstract}

\section*{1. Introduction}

Orlicz spaces are an important concept in analysis and applications (see 19 23, 24). This concept extends the classical concept of \(L^{p}\) Lebesgue spaces for \(p \geq 1\). A convex function \(\Phi(x)\) is used in place of the function \(x^{p}\) appearing in the definition of \(L^{p}\) spaces. This function \(\Phi\) is called a Young function. In addition to \(L^{p}\) spaces, several function spaces can be considered as Orlicz spaces; for example \(L \log ^{+} L\) Zygmund spaces, which are Banach spaces related to Hardy-Littlewood maximal functions. Moreover, Sobolev spaces can be also considered as subspaces of Orlicz spaces (see \(\sqrt{5}\) ). Most of the features of Orlicz spaces have been investigated thoroughly (see 23 , for example), especially, Orlicz spaces determined on measure spaces (see for example 12, 14, 17, 23). In recent years, Orlicz spaces and their weighted cases are examined as Banach algebras over locally compact groups (lcg). Moreover their several properties are also studied (see 1, 20- \(22,27,28\) ).

On the other hand one of the basic problems in harmonic analysis is the description of multipliers. Multipliers have been considered in several contexts, for example Banach algebras and Banach modules theories, partial differential equations, the existence of invariant means, etc. Our aim in this paper is to investigate the affine continuous mappings for the weighted Orlicz space \(L^{\Phi}(\mathbb{A}, \omega)\) over the affine group \(\mathbb{A}\) and study the multiplier problem for \(L^{\Phi}(\mathbb{A}) \cap L^{1}(\mathbb{A})\). The affine

\footnotetext{
2020 Mathematics Subject Classification. 46H25, 43A15, 46E30, 43A22, 43A20.
Keywords. Affine group, affine mapping, multiplier, weighted Orlicz space.
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}
group chosen is a prime example of a nonabelian group on which harmonic analysis and even more applied time-frequency analysis questions are studied (see 8,9 ).

For \(L^{p}\) spaces, in 16, Lau studied the affine mappings \(T\) between the subsets of Lebesgue spaces. In 27, Üster and Öztop studied continuous affine mappings on the subsets of Orlicz spaces. On the other hand the characterization of multipliers for weighted Lebesgue spaces has been given by Gaudry 10. (See also 7.) In 10, Gaudry showed that the multiplier space of \(L^{1}(G, \omega)\) can be characterized by \(M(G, \omega)\). Moreover in 28, Üster characterized the compact mulipliers of \(L^{\Phi}(G, \omega)\). Here \(G\) denotes a lcg. (See Section 2 for notation.)

The paper is organized as follows. In Section 2, we recall some basic definitions and notions on Orlicz and weighted Orlicz spaces. In Section 3, we study continuous affine mappings on subsets of weighted Orlicz space \(L^{\Phi}(\mathbb{A}, \omega)\) and we give a characterization for the multipliers of \(L^{\Phi}(\mathbb{A}) \cap L^{1}(\mathbb{A})\).

\section*{2. Preliminaries}

We start this section by introducing some basic facts for an affine group and essential constructions on it.

Let \(\mathbb{A}:=\left(\mathbb{R}_{+} \times \mathbb{R}, \cdot \mathbb{A}\right)\) be the affine group equipped with the multiplication
\[
\begin{equation*}
(s, t) \cdot \mathbb{A}(x, y)=(s x, s y+t) \tag{1}
\end{equation*}
\]
for \((s, t),(x, y) \in \mathbb{A}\). Note that \((1,0) \cdot \mathbb{A}(s, t)=(s, t) \cdot \mathbb{A}(1,0)=(s, t)\) and \((s, t) \cdot \mathbb{A}\left(s^{-1},-s^{-1} t\right)=\left(s^{-1},-s^{-1} t\right) \cdot \mathbb{A}(s, t)=(1,0)\). Thus \(\mathbb{A}\), endowed with the multiplication (1), becomes a group and this group is called the affine group.

Since a mapping of the real line can be defined by \(F_{s, t}: \mathbb{R} \rightarrow \mathbb{R}\) such that
\[
F_{s, t}(x)=(s, t) \cdot x=s x+t, x \in \mathbb{R}
\]
for any \((s, t) \in \mathbb{A}\), the affine group is also called the \(s x+t\) group. \(F_{s, t}\) is the affine mapping of the real line \(\mathbb{R}\) and this operation is coherent with (1).

We can represent the affine group \(\mathbb{A}\) in matrix form as
\[
\mathbb{A}:=\left\{\left(\begin{array}{cc}
s & t \\
0 & 1
\end{array}\right): s>0, t \in \mathbb{R}\right\} .
\]

The inverse and the identity elements are given by
\[
\left(\begin{array}{cc}
s^{-1} & -s^{-1} t \\
0 & 1
\end{array}\right), \quad I=\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
\]

The operations of the inversion and multiplication are continuous in the product topology. Thus the affine group \(\mathbb{A}\) is a locally compact group and
\[
\begin{aligned}
& d \nu(x, y)=\frac{d x}{x^{2}} d y \\
& d \mu(x, y)=\frac{d x}{x} d y
\end{aligned}
\]
are the left and right Haar measures, respectively (for more details see 13 ). Now since
\[
d \nu(x, y)=\frac{d x}{x^{2}} d y=\frac{1}{x} d \mu(x, y)
\]
the affine group is not unimodular. The modular function on the affine group is \(\Delta(x, y)=x^{-1}\).

Throughout this work we use the right Haar measure \(d \mu\) on \(\mathbb{A}\).
Let \(f: \mathbb{A} \rightarrow \mathbb{C}\) and \((s, t) \in \mathbb{A}\). We use \(L_{(s, t)}\) for the left translation and \(R_{(s, t)}\) for the right translation given by
\[
\left(L_{(s, t)} f\right)(x, y):=f\left((s, t)^{-1} \cdot \mathbb{A}_{\mathbb{A}}(x, y)\right) \quad \text { and } \quad\left(R_{(s, t)} f\right)(x, y):=f\left((x, y) \cdot \mathbb{A}(s, t)^{-1}\right)
\]

Next we give some notions regarding Orlicz spaces, weighted Orlicz spaces and Young functions. Our main references are 12 and 23.
Definition 1. A function \(\Phi:[0, \infty) \rightarrow[0, \infty]\) is called a Young function if \(\Phi\) is convex, \(\Phi(0)=0\) and \(\lim _{t \rightarrow \infty} \Phi(t)=+\infty\).

For a Young function \(\Phi\), its conjugate function \(\Psi\) is given by
\[
\Psi(t)=\sup \{s t-\Phi(s): s \geq 0\} \quad(t \geq 0)
\]

The pair \((\Phi, \Psi)\) of Young functions \(\Phi, \Psi\) is said to be (Young) conjugate and we have
\[
\begin{equation*}
s t \leq \Phi(s)+\Psi(t) \quad(\forall s, t \geq 0) \tag{2}
\end{equation*}
\]

In this paper we only consider the real-valued Young functions. Clearly \(\Phi\) is continuous and \(\lim _{t \rightarrow \infty} \Phi(t)=\infty\). Note that the continuity of \(\Phi\) may not imply the continuity of \(\Psi\).

Let us recall the following facts about Orlicz spaces. Let \((\Phi, \Psi)\) be conjugate Young functions. Then the Orlicz space \(L^{\Phi}(\mathbb{A})\) is defined to be
\[
L^{\Phi}(\mathbb{A})=\left\{f: \mathbb{A} \rightarrow \mathbb{C}: \int_{\mathbb{A}} \Phi(\alpha|f(x, y)|) \frac{d x}{x} d y<\infty \text { for some } \alpha>0\right\}
\]

Here \(f\) and \(g\) in \(L^{\Phi}(\mathbb{A})\) are equivalent if \(f=g\) a.e. Recall that an Orlicz space is a Banach space with respect to (Orlicz) norm which is defined by
\[
\|f\|_{\Phi}=\sup \left\{\int_{\mathbb{A}}|f(x, y) \nu(x, y)| \frac{d x}{x} d y: \int_{\mathbb{A}} \Psi(|\nu(x, y)|) \frac{d x}{x} d y \leq 1\right\}
\]
for \(f \in L^{\Phi}(\mathbb{A})\). Here \((\Phi, \Psi)\) are conjugate Young functions.
Another norm on an Orlicz space is the Luxemburg norm \(N_{\Phi}(f)\) defined by
\[
N_{\Phi}(f)=\inf \left\{\lambda>0: \int_{\mathbb{A}} \Phi\left(\frac{|f(x, y)|}{\lambda}\right) \frac{d x}{x} d y \leq 1\right\}
\]

Note that the Orlicz and Luxemburg norms are equivalent; that is,
\[
N_{\Phi}(\cdot) \leq\|\cdot\|_{\Phi} \leq 2 N_{\Phi}(\cdot)
\]

We shall use the following definition in the last section. In 4 and 29, the main motivation to use this definition is to estimate the norm of the dilation operator. Here we use a result of Lemma 3.3 given in 29 .

Given \(\gamma>0\) one can define
\[
N_{\Phi, \gamma}(f):=\inf \left\{\lambda>0: \int_{\mathbb{A}} \Phi\left(\frac{|f(x, y)|}{\lambda}\right) \frac{d x}{x} d y \leq \gamma\right\}
\]

Here \(N_{\Phi, 1}=N_{\Phi}\) and these norms are equivalent on \(L^{\Phi}(\mathbb{A})\) :
\[
\frac{\gamma_{1}}{\gamma_{2}} N_{\Phi, \gamma_{1}}(f) \leq N_{\Phi, \gamma_{2}}(f) \leq N_{\Phi, \gamma_{1}}(f)
\]
for \(0<\gamma_{1} \leq \gamma_{2}\).
For Orlicz spaces an important notion is the \(\Delta_{2}\)-condition. Let us recall the following definition.
Definition 2. Let \(\Phi:[0, \infty) \rightarrow[0, \infty]\) be a Young function. Then \(\Phi\) is said to satisfy \(\Delta_{2}\)-condition (globally), if
\[
\Phi(2 x) \leq M \Phi(x) \quad(x \geq 0)
\]
for some absolute constant \(M>0\).
Note that if \(\Phi \in \Delta_{2}\), then \(L^{\Phi}(\mathbb{A})^{*} \cong L^{\Psi}(\mathbb{A})\), here \({ }^{*}\) denotes the dual 23 , Corollary 3.4.5]. Moreover if \(\Psi \in \Delta_{2}\), then \(L^{\Phi}(\mathbb{A})\) is a reflexive Banach space (see 1423 for more general cases.)

On the other hand, the weighted Orlicz space \(L^{\Phi}(G, \omega)\) is defined by Osançlol and Öztop in 20 over a lcg \(G\) and they consider the Banach algebra structure for \(L^{\Phi}(G, \omega)\).

A weight function \(\omega\) is a positive, locally integrable function on \(\mathbb{A}\). In this paper we assume that \(\omega\) is continuous (see 25, Section 3.7]). The space \(L^{\Phi}(\mathbb{A}, \omega)\) is defined by \(\left\{f: f \omega \in L^{\Phi}(\mathbb{A})\right\}\). We also set
\[
\begin{equation*}
N_{\Phi}^{\omega}(f)=N_{\Phi}(f \omega) \tag{3}
\end{equation*}
\]
for \(f \in L^{\Phi}(\mathbb{A}, \omega)\). Then \(N_{\Phi}^{\omega}(\cdot)\) defines a norm on \(L^{\Phi}(\mathbb{A}, \omega)\) and \(L^{\Phi}(\mathbb{A}, \omega)\) is a Banach space with respect to this norm. Moreover, \(L^{\Psi}\left(\mathbb{A}, \omega^{-1}\right)\) is the dual space of \(\left(L^{\Phi}(\mathbb{A}, \omega), N_{\Phi}^{\omega}(\cdot)\right)\) if \(\Phi\) fulfills the \(\Delta_{2}\)-condition. Here the duality is given by
\[
\langle f, h\rangle=\int_{\mathbb{A}} f(x, y) h(x, y) \frac{d x}{x} d y \quad\left(f \in L^{\Phi}(\mathbb{A}, \omega), h \in L^{\Psi}\left(\mathbb{A}, \omega^{-1}\right)\right)
\]
where \((\Phi, \Psi)\) are conjugate Young functions and the space \(L^{\Psi}\left(\mathbb{A}, \omega^{-1}\right)\) is endowed with the norm \(N_{\Psi}^{\omega^{-1}}(f)=N_{\Psi}\left(\frac{f}{\omega}\right)\). So if \(\Phi, \Psi\) fulfill the \(\Delta_{2}\)-condition then \(L^{\Phi}(\mathbb{A}, \omega)\) is a reflexive Banach space (for the general case see 20 ).

For \(\Phi(x)=\frac{x^{p}}{p}, 1<p<\infty\), the conjugate Young function is \(\Psi(y)=\frac{y^{q}}{q}\), where \(\frac{1}{p}+\frac{1}{q}=1\). Then \(L^{\Phi}(\mathbb{A}, \omega)\) and its norm are equal to the Lebesgue space \(L^{p}(\mathbb{A}, \omega)\)
and its norm. For \(p=1\) and \(\Phi(x)=x\) the conjugate Young function is
\[
\Psi(y)= \begin{cases}0, & 0 \leq y \leq 1 \\ \infty, & \text { otherwise }\end{cases}
\]
and we have \(L^{\Phi}(\mathbb{A}, \omega)=L^{1}(\mathbb{A}, \omega)\). Note that for \(p=1\), the Banach algebra \(L^{1}(\mathbb{A}, \omega)\) always has a bounded approximate identity.

As usual, \(M(\mathbb{A}, \omega)\) is the set of all complex bounded regular Borel measures \(\lambda\) on \(\mathbb{A}\) with
\[
\|\lambda\|_{\omega}=\int_{\mathbb{A}} \omega(s, t) d \lambda(s, t)<\infty
\]

We denote the space of all continuous functions \(f\) on \(\mathbb{A}\) vanishing at infinity by \(C^{0}\left(\mathbb{A}, \omega^{-1}\right)\) with the norm \(\|f\|_{\infty, \omega^{-1}}=\left\|\frac{f}{\omega}\right\|_{\infty}\). Then \(M(\mathbb{A}, \omega)\) is realized as \(\left(C^{0}\left(\mathbb{A}, \omega^{-1}\right)\right)^{*}\) by
\[
\langle\lambda, f\rangle=\int_{\mathbb{A}} f(x, y) d \lambda(x, y)
\]
(for the general case see 11 ). If \(\lambda \in M(\mathbb{A}, \omega)\) and \(f \in L^{\Phi}(\mathbb{A}, \omega)\) the convolution of \(\lambda\) and \(f\) is defined by
\[
(\lambda * f)(x, y)=\int_{\mathbb{A}} f\left((s, t)^{-1} \cdot \mathbb{A}(x, y)\right) d \lambda(s, t)
\]

Moreover if \(f, g\) are measurable functions on \(\mathbb{A}\) the convolution of \(f\) and \(g\) is defined by
\[
(f * g)(x, y)=\int_{\mathbb{A}} f(s, t) g\left((s, t)^{-1} \cdot \mathbb{A}(x, y)\right) \frac{d s}{s} d t \quad((x, y) \in \mathbb{A})
\]

For each \((s, t) \in \mathbb{A}\), let \(\delta_{(s, t)}(E)=1_{E}(s, t)\), where \(1_{E}\) is the characteristic function of \(E \subseteq \mathbb{A}\). Then
\[
\left(\delta_{(s, t)} * f\right)(x, y)=f\left((s, t)^{-1} \cdot \mathbb{A}(x, y)\right)=L_{(s, t)} f(x, y) \quad((s, t) \in \mathbb{A})
\]
where \(L_{(s, t))^{-1}}\) is the left translation operator. For a function \(f\) on \(\mathbb{A}\), we use \(\tilde{f}\) defined by \(\widetilde{f}(x, y)=f\left((x, y)^{-1}\right)\) for each \((x, y) \in \mathbb{A}\).

Throughout the paper we study \(L^{\Phi}(\mathbb{A}, \omega)\) with the weight \(\omega\) and the \(\Delta_{2}\)-condition on a Young function \(\Phi\).

\section*{3. Main Results}

In this section we characterize the affine continuous mappings for \(L^{\Phi}(\mathbb{A}, \omega)\) over the affine gorup \(\mathbb{A}\) and we study the multiplier problem for the space \(L^{\Phi}(\mathbb{A}, \omega) \cap\) \(L^{1}(\mathbb{A}, \omega)\). Let us first give the following definitions.

Definition 3. Let \(C \subseteq L^{\Phi}(\mathbb{A}, \omega)\). Then \(C\) is called left invariant if \(L_{(x, y)} f \in C\) for each \(f \in C\) and \((x, y) \in \mathbb{A}\).

Notice that for \(f \in L^{\Phi}(\mathbb{A}, \omega)\) and \((x, y) \in \mathbb{A}\) we have \(L_{(x, y)} f \in L^{\Phi}(\mathbb{A}, \omega)\) and \(N_{\Phi}^{\omega}\left(L_{(x, y)} f\right) \leq \omega(x, y) N_{\Phi}^{\omega}(f)\) (for the general lcgs see 20, Lemma 2.3]).

Definition 4. Let \(X\) and \(Y\) be normed spaces and \(C, D\) be convex subsets of \(X\) and \(Y\) respectively. Then a mapping \(f: C \rightarrow D\) is called affine if
\[
f(\alpha x+(1-\alpha) y)=\alpha f(x)+(1-\alpha) f(y)
\]
for each \(x, y \in C\) and \(\alpha \in[0,1]\).
For the subset \(K\) of \(L^{\Phi}(\mathbb{A}, \omega)\), we use co \(K\) for the convex hull of \(K\). In addition to the norm topology on \(L^{\Phi}(\mathbb{A}, \omega)\), we will take the weak topology \(w\) and the weak* topology \(w^{*}\) for the pair \(\left(L^{\Phi}(\mathbb{A}, \omega), L^{\Phi}(\mathbb{A}, \omega)^{*}\right)\) where \((\Phi, \Psi)\) is a conjugate pair.

Moreover, we make use of the following subsets of \(M(\mathbb{A}, \omega)\) :
(i) \(P(\mathbb{A}, \omega)=\left\{\mu \in M(\mathbb{A}, \omega):\|\mu\|_{\omega}=1\right.\) and \(\left.\mu \geq 0\right\}\),
(ii) \(P_{1}(\mathbb{A}, \omega)=\left\{h \in L^{1}(\mathbb{A}, \omega):\|h\|_{1, \omega}=1\right.\) and \(\left.h \geq 0\right\}\),
(iii) \(E(\mathbb{A}, \omega)=\left\{\frac{\delta_{(x, y)}}{\omega(x, y)}:(x, y) \in \mathbb{A}\right\}\).

We omit the proof of the following Lemma which appears in 28 for general locally compact abelian groups. One can get the same result for nonabelian groups in a similar way.

Lemma 1. We have \(P(\mathbb{A}, \omega)={\overline{P_{1}(\mathbb{A}, \omega)}}^{w^{*}}=\overline{\operatorname{co} E(\mathbb{A}, \omega)}^{w^{*}}\). Here \(\bar{\sigma}^{w^{*}}\) indicates weak* closure.

Lemma 2. The following are true.
(i) Let \(f \in L^{\Phi}(\mathbb{A}, \omega)\). Then the mapping \(\mu \mapsto \mu * f\) is continuous from \(\left(M(\mathbb{A}, \omega), w^{*}\right)\) to \(\left(L^{\Phi}(\mathbb{A}, \omega), w\right)\).
(ii) Let \(f \in L^{1}(\mathbb{A}, \omega)\). Then the mapping \(h \mapsto f * h\) is continuous from \(\left(L^{\Phi}(\mathbb{A}, \omega), w\right)\) to \(\left(L^{\Phi}(\mathbb{A}, \omega), w\right)\).
Proof. (i) Let \(\left\{\mu_{\alpha}\right\}_{\alpha} \subseteq M(\mathbb{A}, \omega)\) be a net that is weak* convergent to \(\mu,(\Phi, \Psi)\) a conjugate Young pair and \(f \in L^{\Phi}(\mathbb{A}, \omega)\). Since \(L^{\Phi}(\mathbb{A}, \omega)\) is \(M(\mathbb{A}, \omega)\)-module the mapping \(\mu * f\) is well defined (see 20\()\). Let \(T \in\left(L^{\Phi}(\mathbb{A}, \omega)\right)^{*}\), so there exists \(g \in L^{\Psi}\left(\mathbb{A}, \omega^{-1}\right)\) such that
\[
T(f)=\int_{\mathbb{A}} f(x, y) g(x, y) \frac{d x}{x} d y=\langle f, g\rangle
\]

Thus we obtain that
\[
\begin{aligned}
T\left(\mu_{\alpha} * f\right) & =\left\langle\mu_{\alpha} * f, g\right\rangle \\
& =\int_{\mathbb{A}}\left(\mu_{\alpha} * f\right)(x, y) g(x, y) \frac{d x}{x} d y
\end{aligned}
\]
\[
\begin{aligned}
& =\int_{\mathbb{A}} \int_{\mathbb{A}} f\left((s, t)^{-1} \cdot \mathbb{A}\right. \\
& =\int_{\mathbb{A}} \int_{\mathbb{A}} \tilde{f}((x, y)) d \mu_{\alpha}(s, t) g(x, y) \frac{d x}{x} d y \\
& =\int_{\mathbb{A}}(g * \tilde{f})(s, t) d \mu_{\alpha}(s, t) \\
& =\left\langle g * \tilde{f}, \mu_{\alpha}\right\rangle .
\end{aligned}
\]

Since \(f \in L^{\Phi}(\mathbb{A}, \omega)\) and \(g \in L^{\Psi}\left(\mathbb{A}, \omega^{-1}\right)\), we have \(g * \tilde{f} \in C_{0}\left(\mathbb{A}, \omega^{-1}\right)\) (for the general case see 20.) This implies that \(T\left(\mu_{\alpha} * f\right)=\left\langle g * \tilde{f}, \mu_{\alpha}\right\rangle \rightarrow\langle g * \tilde{f}, \mu\rangle=\langle\mu * f, g\rangle=\) \(T(\mu * f)\), i.e., \(\mu_{\alpha} * f\) weakly converges to \(\mu * f\) in \(L^{\Phi}(\mathbb{A}, \omega)\).
(ii) Let \(\left\{h_{\alpha}\right\}_{\alpha} \subseteq L^{\Phi}(\mathbb{A}, \omega)\) be a net that is weakly convergent to \(h\) and \(f \in\) \(L^{1}(\mathbb{A}, \omega)\). We have \(\lim _{\alpha}\left\langle h_{\alpha}, g\right\rangle=\langle h, g\rangle\) for all \(g \in\left(L^{\Phi}(\mathbb{A}, \omega)\right)^{*}\). Thus we obtain that
\[
\begin{aligned}
\left\langle h_{\alpha} * f, g\right\rangle & =\int_{\mathbb{A}}\left(h_{\alpha} * f\right)(x, y) g(x, y) \frac{d x}{x} d y \\
& =\int_{\mathbb{A}} \int_{\mathbb{A}} h_{\alpha}(s, t) f\left((s, t)^{-1} \cdot \mathbb{A}(x, y)\right) g(x, y) \frac{d s}{s} d t \frac{d x}{x} d y \\
& =\int_{\mathbb{A}} \int_{\mathbb{A}} h_{\alpha}(s, t) \tilde{f}\left((x, y)^{-1} \cdot{ }_{\mathbb{A}}(s, t)\right) g(x, y) \frac{d s}{s} d t \frac{d x}{x} d y \\
& =\int_{\mathbb{A}} h_{\alpha}(s, t)(g * \tilde{f})(s, t) \frac{d s}{s} d t \\
& =\left\langle h_{\alpha}, g * \tilde{f}\right\rangle
\end{aligned}
\]

This gives that \(\left\langle h_{\alpha} * f, g\right\rangle=\left\langle h_{\alpha}, g * \tilde{f}\right\rangle \rightarrow\langle h, g * \tilde{f}\rangle=\langle h * f, g\rangle\).
Theorem 1. Let \(C, D\) be convex, closed, left invariant subsets of \(L^{\Phi_{2}}(\mathbb{A}, \omega)\) and \(L^{\Phi_{1}}(\mathbb{A}, \omega)\) respectively. If \(T: C \rightarrow D\) is a continuous and affine mapping then the following are equivalent.
(i) \(T\left(L_{(x, y)} f\right)=L_{(x, y)}(T f)\) for each \((x, y) \in \mathbb{A}\) and \(f \in C\).
(ii) \(T(\nu * f)=\nu * T(f)\) for each \(\nu \in P_{1}(\mathbb{A}, \omega)\) and \(f \in C\).

Proof. (i \(\Rightarrow\) ii) Let \(f \in C\) and assume that \(T\left(L_{(x, y)} f\right)=L_{(x, y)}(T f)\) for each \((x, y) \in \mathbb{A}\) and \(\nu \in P_{1}(\mathbb{A}, \omega)\). Using Lemma 1, there exists a net \(\left\{\nu_{\alpha}\right\}_{\alpha}\) in \(\operatorname{co} E(\mathbb{A}, \omega)\), \(\nu_{\alpha}=\sum_{i=1}^{n_{\alpha}} \lambda_{i}{ }^{\alpha} \frac{\delta_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}^{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}}{}\) and \(\nu_{\alpha}\) weak \(^{*}\) converges to \(\nu\). Then by Lemma \(2,\left\{\nu_{\alpha} * f\right\}_{\alpha}\)
weakly converges to \(\nu * f\) for each \(f \in C\). Thus we have
\[
\nu_{\alpha} * f=\left(\sum_{i=1}^{n_{\alpha}} \lambda_{i}^{\alpha} \frac{\delta_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}}{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}\right) * f=\sum_{i=1}^{n_{\alpha}} \frac{\lambda_{i}^{\alpha}}{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)} L_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)^{-1}} f .
\]

As \(C\) is convex and left invariant, the net \(\left\{\nu_{\alpha} * f\right\}_{\alpha}\) is contained in \(C\). Now using Lemma 2 it follows that \(\nu * f \in C\).

On the other hand since \(C\) and \(D\) are convex and closed they are weakly closed. Moreover since \(T\) is continuous and affine \(T\) is weakly continuous when \(C\) and \(D\) have their respective weak topologies (see 6, 26). Then we get that
\[
\begin{aligned}
& T(\nu * f)=\lim _{\alpha} T\left(\nu_{\alpha} * f\right) \\
&=\lim _{\alpha} T\left(\left(\sum_{i=1}^{n_{\alpha}} \lambda_{i}^{\alpha} \frac{\left.\left.\delta_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}^{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}\right) * f\right)}{}\right.\right. \\
&=\lim _{\alpha} T\left(\sum_{i=1}^{n_{\alpha}} \frac{\lambda_{i}^{\alpha}}{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}\left(\delta_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)} * f\right)\right. \\
&=\lim _{\alpha} T\left(\sum_{i=1}^{n_{\alpha}} \frac{\lambda_{i}^{\alpha}}{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}\left(L_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)^{-1}} f\right)\right) \\
&=\lim _{\alpha} \sum_{i=1}^{n_{\alpha}} \frac{\lambda_{i}^{\alpha}}{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)} T\left(L_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)^{-1}} f\right) \\
&=\lim _{\alpha} \sum_{i=1}^{n_{\alpha}} \frac{\lambda_{i}^{\alpha}}{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)} L_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)^{-1}} T(f) \\
&=\lim _{\alpha}\left(\sum_{i=1}^{n_{\alpha}} \lambda_{i}^{\alpha} \frac{\left.\delta_{\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}^{\omega\left(s_{i}^{\alpha}, t_{i}^{\alpha}\right)}\right) * T(f)}{}\right. \\
&=\lim _{\alpha} \nu_{\alpha} * T(f) \\
&=\nu * T(f) .
\end{aligned}
\]
(ii \(\Rightarrow\) i) Conversely let \((x, y) \in \mathbb{A}\). Using Lemma 1 there exists a net \(\left\{\nu_{\alpha}\right\}_{\alpha} \subseteq\) \(P_{1}(\mathbb{A}, \omega)\) such that \(\left\{\nu_{\alpha}\right\}_{\alpha}\) converges to \(\delta_{(x, y)^{-1}}\) in the weak* topology. If \(T(\nu * f)=\) \(\nu * T(f)\) for each \(\nu \in P_{1}(\mathbb{A}, \omega)\) and \(f \in C\) we have that
\[
\begin{aligned}
T\left(L_{(x, y)} f\right) & =T\left(\delta_{(x, y)^{-1}} * f\right) \\
& =\lim _{\alpha} T\left(\nu_{\alpha} * f\right) \\
& =\lim _{\alpha} \nu_{\alpha} * T(f) \\
& =\delta_{(x, y)^{-1}} * T(f) \\
& =L_{(x, y)} T(f) .
\end{aligned}
\]

This completes the proof.
Theorem 2. Let \(B\) be a weakly compact, bounded, left invariant, closed subset of \(L^{\Phi}(\mathbb{A}, \omega)\) and \(T\) be a continuous affine mapping from \(P_{1}(\mathbb{A}, \omega)\) to \(B\). Then \(T\) commutes with all left translations if and only if there exists an \(f \in B\) such that \(T(g)=g * f\) for each \(g \in P_{1}(\mathbb{A}, \omega)\).

Proof. Let \((x, y) \in \mathbb{A}\) and assume that \(T\left(L_{(x, y)} g\right)=L_{(x, y)}(T g)\) for each \(g \in\) \(P_{1}(\mathbb{A}, \omega)\). Using Theorem 1 we have \(T(k * g)=k * T(g)\) for \(k, g \in P_{1}(\mathbb{A}, \omega)\). Let \(\left\{u_{\alpha}\right\}_{\alpha} \subseteq P_{1}(\mathbb{A}, \omega)\) be a bounded approximate identity for \(L^{1}(\mathbb{A}, \omega)\). Since \(B\) is weakly compact and \(T\left(u_{\alpha}\right) \in B\) is bounded, there exists \(f \in B\) such that \(\left\{T\left(u_{\alpha}\right)\right\}_{\alpha}\) converges to \(f\) weakly. Thus
\[
\begin{aligned}
T(g) & =\lim _{\alpha} T\left(g * u_{\alpha}\right) \\
& =\lim _{\alpha} g * T\left(u_{\alpha}\right) \\
& =g * f
\end{aligned}
\]
and the result follows.
For the converse let \((x, y) \in \mathbb{A}\) and assume that \(f \in B\) such that \(T(g)=g * f\) for all \(g \in P_{1}(\mathbb{A}, \omega)\). Then
\[
\begin{aligned}
L_{(x, y)} T(g) & =L_{(x, y)}(g * f) \\
& =\delta_{(x, y)^{-1}} *(g * f) \\
& =\left(\delta_{(x, y)^{-1}} * g\right) * f \\
& =L_{(x, y)} g * f \\
& =T\left(L_{(x, y)} g\right)
\end{aligned}
\]
which gives the required result.
Now our purpose is to obtain a characterization for the multipliers of \(L^{\Phi}(\mathbb{A}) \cap\) \(L^{1}(\mathbb{A})\). We observe that the following result does not work for the weighted case and we give the result for the unweighted case.

We start with the definition of the left multiplier of \(L^{\Phi}(\mathbb{A})\).
Definition 5. Let \(T\) be a bounded linear operator from \(L^{\Phi_{1}}(\mathbb{A})\) to \(L^{\Phi_{2}}(\mathbb{A})\). Then \(T\) is said to be a left multiplier for \(\left(L^{\Phi_{2}}(\mathbb{A}), L^{\Phi_{1}}(\mathbb{A})\right)\) if \(T\left(L_{(x, y)} f\right)=L_{(x, y)}(T f)\) for all \(f \in L^{\Phi_{2}}(\mathbb{A})\) and \((x, y) \in \mathbb{A}\). We write \(\mathcal{M}\left(L^{\Phi_{2}}(\mathbb{A}), L^{\Phi_{1}}(\mathbb{A})\right)\) for the set of left multipliers of \(\left(L^{\Phi_{2}}(\mathbb{A}), L^{\Phi_{1}}(\mathbb{A})\right)\).

Remark 1. Observe that the normed space \(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\) is a Banach space with the norm
\[
\mid\|f\|\|=\| f \|_{1}+N_{\Phi}(f)
\]
and dense in \(L^{1}(\mathbb{A})\).

The following lemma is important to us for our last result (for the proof see 29 , Lemma 3.3].)
Lemma 3. Let \(\Phi\) be a Young function satisfying the \(\Delta_{2}\) condition. If \(f \in L^{\Phi}(\mathbb{A})\) then \(\lim _{(a, b) \rightarrow(+\infty,+\infty)} N_{\Phi}\left(f+L_{(a, b)} f\right)=N_{\Phi, \frac{1}{2}}(f)\).

Now we have the tools to give a characterization of the multipliers of \(L^{\Phi}(\mathbb{A}) \cap\) \(L^{1}(\mathbb{A})\).

Theorem 3. Let \(T: L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}) \rightarrow L^{1}(\mathbb{A})\) be a linear mapping. Then the following are equivalent.
(i) \(T \in \mathcal{M}\left(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}), L^{1}(\mathbb{A})\right)\).
(ii) There exists a unique measure \(\mu \in M(\mathbb{A})\) such that \(T f=\mu * f\) for each \(f \in L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\).
Furthermore the correspondence between \(T\) and \(\mu\) defines an isometric isomorphism of \(\mathcal{M}\left(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}), L^{1}(\mathbb{A})\right)\) onto \(M(\mathbb{A})\).

Proof. Assume that \(T \in \mathcal{M}\left(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}), L^{1}(\mathbb{A})\right)\). Then for each \(f \in L^{1}(\mathbb{A}) \cap\) \(L^{\Phi}(\mathbb{A})\) we obtain that
\[
\begin{equation*}
\|T f\|_{1} \leq\|T\|\left(\|f\|_{1}+N_{\Phi}(f)\right) \tag{4}
\end{equation*}
\]

By Lemma 3 we have \(\lim _{(s, t) \rightarrow(\infty, \infty)} N_{\Phi}\left(f+L_{(s, t)} f\right)=N_{\Phi, \frac{1}{2}}(f)\). Using this fact together with (4) we have that
\[
\begin{aligned}
2\|T f\|_{1} & =\lim _{(s, t) \rightarrow(\infty, \infty)}\left\|T f+L_{(s, t)} T f\right\|_{1} \\
& =\lim _{(s, t) \rightarrow(\infty, \infty)}\left\|T\left(f+L_{(s, t)} f\right)\right\|_{1} \\
& \leq \lim _{(s, t) \rightarrow(\infty, \infty)}\|T\|\left(\left\|f+L_{(s, t)} f\right\|_{1}+N_{\Phi}\left(f+L_{(s, t)} f\right)\right) \\
& =\|T\|\left(2\|f\|_{1}+N_{\Phi, \frac{1}{2}}(f)\right)
\end{aligned}
\]
for each \(f \in L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\). Therefore we obtain
\[
\|T f\|_{1} \leq\|T\|\left(\|f\|_{1}+2^{-1} N_{\Phi, \frac{1}{2}}(f)\right)
\]

Applying this step \(n\) times we obtain
\[
\|T f\|_{1} \leq\|T\|\left(\|f\|_{1}+2^{-n} N_{\Phi, \frac{1}{2}}(f)\right)
\]
for \(f \in L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\). Since \(\lim _{n \rightarrow \infty} 2^{-n}=0\) we deduce that \(\|T f\|_{1} \leq\|T\|\|f\|_{1}\).
Thus \(T\) defines a linear continuous mapping from \(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\) to \(L^{1}(\mathbb{A})\) commuting with left translations. Moreover since \(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\) is dense in \(L^{1}(\mathbb{A}), T\) determines a unique \(\operatorname{map} S \in \mathcal{M}\left(L^{1}(\mathbb{A})\right)\) and \(\|S\| \leq\|T\|\). Moreover there exists a unique \(\mu \in M(\mathbb{A})\) such that \(S f=\mu * f\) for each \(f \in L^{1}(\mathbb{A})\) and \(\|\mu\|=\|S\|\) (see 30 ). Therefore \(T f=\mu * f\) for each \(f \in L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\) and \(\|\mu\| \leq\|T\|\).

Conversely, if \(\mu \in M(\mathbb{A})\) and \(T f=\mu * f\) for each \(f \in L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A})\) we obtain
\[
\|T f\|_{1}=\|\mu * f\|_{1} \leq\|\mu\|\|f\|_{1} \leq\|\mu\|\|f \mid\| .
\]

Therefore \(T \in \mathcal{M}\left(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}), L^{1}(\mathbb{A})\right)\) and \(\|T\| \leq\|\mu\|\).
This gives to equivalence of (i) and (ii).
It is clear that the correspondence between \(T\) and \(\mu\) defines an isometric isomor\(\operatorname{phism}\) from \(\mathcal{M}\left(L^{1}(\mathbb{A}) \cap L^{\Phi}(\mathbb{A}), L^{1}(\mathbb{A})\right)\) onto \(M(\mathbb{A})\).

Declaration of Competing Interests The author declares that there are no conflicts of interest regarding the publication of this paper.
Acknowledgements The author is grateful to an anonymous referee for a careful reading of the manuscript and for helpful comments.

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TRANSLATION SURFACES GENERATING WITH SOME PARTNER CURVE
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\begin{abstract}
In this article, generating curves of translation surfaces are paired with some special curve pairs. With the results obtained from these pairings, the developable and minimal translation surfaces are characterized. In addition, the surface curvatures of the translation surface are obtained. For a better understanding of the results, examples are given and their drawings are made with the help of Mathematica.
\end{abstract}

\section*{1. Introduction}

The main purpose of differential geometry is to understand and characterize the mathematical properties of any geometric object defined in space. The most important of these objects are curves and surfaces. Researchers working on this subject often have to characterize the curve and the surface in a certain way in order to understand it. One of the most important ways to characterize the curve is to use Frenet vectors. For example, Bertrand pairs of curves were characterized by J. Bertrand in 1850 as curves whose reciprocal normal vectors are linearly dependent 1. Similarly, the Mannheim curve pairs were characterized by the normal vector of one of the curves and the binormal vector of the other as linearly dependent by A. Mannheim in 1878 [2. In addition, the involute-evolute curve pairs are characterized as curve pairs whose mutual tangent vectors are perpendicular 3 .

The study of surfaces is one of the most captivating subjects in the field of differential geometry. Consequently, researchers have extensively investigated various types of surfaces \(4-6\). Much like curves, researchers endeavor to characterize

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2020 Mathematics Subject Classification. 53A04, 53A05.
Keywords. Translation surface, Bertrand partner curve, Mannheim partner curve, involuteevolute curves.
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surfaces. Moreover, another significant aspect that piques researchers' interest is whether a surface is developable or minimal 7.8. One of the interesting surfaces in Euclidean space is the translation surface produced by the two curves. The general form of translation surface is the surface that can be generated from two arbitrary space curves by translating either of them parallel to itself. In such a way that each of its points describes a curve that is a translation of the other curve. A generalized type of translation surface parameterized by
\[
\begin{equation*}
\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v) \tag{1}
\end{equation*}
\]
where \(\mathbf{x}: I \subset \mathbb{R} \rightarrow \mathbb{E}^{3}\) and \(\mathbf{y}: J \subset \mathbb{R} \rightarrow \mathbb{E}^{3}\) are arbitrary generating curves of \(\chi\) according to the parameters \(u\) and \(v\) (may be the arc-length parameters), respectively. Let \(\left\{\mathbf{t}_{\mathbf{x}}, \mathbf{n}_{\mathbf{x}}, \mathbf{b}_{\mathbf{x}}\right\}\) be the Frenet frame field of \(\mathbf{x}\) with curvature \(\kappa_{\mathbf{x}}\) and torsion \(\tau_{\mathbf{x}}\). Also, let \(\left\{\mathbf{t}_{\mathbf{y}}, \mathbf{n}_{\mathbf{y}}, \mathbf{b}_{\mathbf{y}}\right\}\) be the Frenet frame field of \(\mathbf{y}\) with curvature \(\kappa_{\mathbf{y}}\) and torsion \(\tau_{\mathbf{y}}\). A translation surface has the property that the translations of a parametric curve \(u=c\) by \(\mathbf{y}(v)\) remain in \(\chi\) (similarly for the parametric curves \(v=c\) ) 9-11. Translation surfaces are the basic modeling surfaces commonly used in computer aided geometric design and geometric modeling 12. Also, translation surfaces are common in descriptive geometry and architecture because they can be easily modeled 13, 14. Many studies are carried out on translation surfaces so far: L. Verstraelen et al. have studied minimal translation surfaces in n-dimensional Euclidean spaces 15. H. Liu has studied Gaussian curvature and mean curvature of translation surfaces in 3-dimensional space 16. D. W. Yoon has studied the differential geometric properties of translation surfaces by applying the Laplace operator to the Gauss transform 17. Additionally, numerous studies have been conducted on translation surfaces \(18-22\).

In this study, generating curves of translation surfaces are associated with some special curve pairs. The article investigates the conditions necessary for these translation surfaces to be both developable and minimal surface, while also characterizing the conditions that make this possible.

\section*{2. Preliminaries}

In this section, for parametrized curves and surface elements some basics definitions and theorems are given.

A regular naturally parametrization of class \(C^{k}\), with \(k \geq 1\) of a curve in \(\mathbb{R}^{3}\) is a vector function \(\mathbf{x}: I \subset \mathbb{R} \rightarrow \mathbb{E}^{3}, s \mapsto \mathbf{x}(s)=\left(\mathbf{x}_{1}(s), \mathbf{x}_{2}(s), \mathbf{x}_{3}(s)\right)\) defined on an interval \(I\) which satisfies \(\mathbf{x}\) is of class \(C^{k}\) and \(\mathbf{x}^{\prime}(s) \neq 0\) for all \(s \in I\). A curve \(\mathbf{x}\) is continuously differentiable if \(\mathbf{x}^{\prime}(s)\) exists for all \(s \in I\) and the derivative \(\mathbf{x}^{\prime}(s)\) is a continuous function; thinking dynamically, the vector \(\mathbf{x}^{\prime}(s)\) is the velocity of the curve at time \(s\). We call \(\mathbf{x}(s)\) naturally parametrized curve if \(\mathbf{x}_{i}(s)(i=1,2,3)\) is of class \(C^{k}\) and \(\left\|\mathbf{x}^{\prime}(s)\right\|=1\), for each \(s \in I \quad 23\).

Let \(\mathbf{x}(s)\) be biregular, that is, \(\mathbf{x}^{\prime}(s) \times \mathbf{x}^{\prime \prime}(s) \neq 0\), for each \(s \in I\). We consider a trihedron \(\{\mathbf{t}(s), \mathbf{n}(s), \mathbf{b}(s)\}\) along \(\mathbf{x}(s)\), so-called Frenet frame, where 23
\[
\mathbf{t}(s)=\mathbf{x}^{\prime}(s), \quad \mathbf{n}(s)=\frac{\mathbf{t}^{\prime}(s)}{\left\|\mathbf{t}^{\prime}(s)\right\|}, \quad \mathbf{b}(s)=\mathbf{t}(s) \times \mathbf{n}(s)
\]

The curvature \(\kappa\), a non-negative scalar field, is defined by setting \(\kappa(s)=\left\|\mathbf{t}^{\prime}(s)\right\|\) and torsion is defined by setting \(\tau(s)=\left\langle\mathbf{n}^{\prime}(s), \mathbf{b}(s)\right\rangle\). If the naturally parametrized curve \(\mathbf{x}\) has unit speed and strictly positive curvature, then the following equations hold 23
\[
\left[\begin{array}{l}
\mathbf{t}^{\prime} \\
\mathbf{n}^{\prime} \\
\mathbf{b}^{\prime}
\end{array}\right]=\left[\begin{array}{ccc}
0 & \kappa & 0 \\
-\kappa & 0 & \tau \\
0 & -\tau & 0
\end{array}\right]\left[\begin{array}{l}
\mathbf{t} \\
\mathbf{n} \\
\mathbf{b}
\end{array}\right] .
\]
where \(\kappa \neq 0\) for the Frenet frame to be defined.
Let \(\mathbf{x}\) and \(\mathbf{y}\) be naturally parametrized curves in \(\mathbb{E}^{3}\) with parameter \(u\) and \(v\), respectively. Let \(\left\{t_{\mathbf{x}}(u), n_{\mathbf{x}}(u), b_{\mathbf{x}}(u), \kappa_{\mathbf{x}}(u), \tau_{\mathbf{x}}(u)\right\}\) and \(\left\{t_{\mathbf{y}}(v), n_{\mathbf{y}}(v), b_{\mathbf{y}}(v), \kappa_{\mathbf{y}}(v), \tau_{\mathbf{y}}(v)\right\}\) be Frenet elements of \(\mathbf{x}\) and \(\mathbf{y}\), respectively. Some special curve pairs is studied by S. Yuce and A. Sabuncuoglu and the following results are given \(24,25\).

Let's assume that ( \(\mathbf{x}, \mathbf{y}\) ) curve pair is Bertrand curve pair. In this sitation, since the normal vectors of the \(\mathbf{x}\) and \(\mathbf{y}\) have the same direction, they are written as
\[
\begin{align*}
\mathbf{t}_{\mathbf{x}}(u) & =\cos \theta \mathbf{t}_{\mathbf{y}}(v)-\sin \theta \mathbf{b}_{\mathbf{y}}(v)  \tag{2}\\
\mathbf{n}_{\mathbf{x}}(u) & =\mathbf{n}_{\mathbf{y}}(v)  \tag{3}\\
\mathbf{b}_{\mathbf{x}}(u) & =\sin \theta \mathbf{t}_{\mathbf{y}}(v)+\cos \theta \mathbf{b}_{\mathbf{y}}(v) \tag{4}
\end{align*}
\]
and
\[
\begin{align*}
\kappa_{\mathbf{x}}(u) & =\kappa_{\mathbf{y}}(v) \cos \theta+\tau_{\mathbf{y}}(v) \sin \theta  \tag{5}\\
\tau_{\mathbf{x}}(u) & =-\kappa_{\mathbf{y}}(v) \sin \theta+\tau_{\mathbf{y}}(v) \cos \theta \tag{6}
\end{align*}
\]
where \(\theta\) is the constant angle between the mutually tangent vectors.
Let's assume that \((\mathbf{x}, \mathbf{y})\) curve pair is Mannheim curve pair. Since the normal vector of the \(\mathbf{x}\) and binormal vector of the curve \(\mathbf{y}\) have the same direction, they are written as
\[
\begin{align*}
\mathbf{t}_{\mathbf{x}}(u) & =\cos \theta \mathbf{t}_{\mathbf{y}}(v)+\sin \theta \mathbf{n}_{\mathbf{y}}(v)  \tag{7}\\
\mathbf{n}_{\mathbf{x}}(u) & =\mathbf{b}_{\mathbf{y}}(v)  \tag{8}\\
\mathbf{b}_{\mathbf{x}}(u) & =-\sin \theta \mathbf{t}_{\mathbf{y}}(v)+\cos \theta \mathbf{n}_{\mathbf{y}}(v) \tag{9}
\end{align*}
\]
and
\[
\begin{align*}
\kappa_{\mathbf{x}}(u) & =\tau_{\mathbf{y}}(v) \sin \theta \frac{d v}{d u}  \tag{10}\\
\tau_{\mathbf{x}}(u) & =-\tau_{\mathbf{y}}(v) \cos \theta \frac{d v}{d u} \tag{11}
\end{align*}
\]
where \(\theta\) is the constant angle between the mutually tangent vectors.

Let's assume that \((\mathbf{x}, \mathbf{y})\) curve pair be involute-evolute partner curve. Since the mutual tangent vectors of the \(\mathbf{x}\) and \(\mathbf{y}\) curves are perpendicular, the following equations are available
\[
\begin{align*}
\mathbf{t}_{\mathbf{x}}(u) & =\mathbf{n}_{\mathbf{y}}(v)  \tag{12}\\
\mathbf{n}_{\mathbf{x}}(u) & =\cos \theta \mathbf{t}_{\mathbf{y}}(v)+\sin \theta \mathbf{b}_{\mathbf{y}}(v)  \tag{13}\\
\mathbf{b}_{\mathbf{x}}(u) & =-\sin \theta \mathbf{t}_{\mathbf{y}}(v)+\cos \theta \mathbf{b}_{\mathbf{y}}(v) \tag{14}
\end{align*}
\]
where \(\theta\) is the constant angle between \(\mathbf{t}_{\mathbf{x}}\) and \(\mathbf{n}_{\mathbf{y}}\), and
\[
\begin{equation*}
\kappa_{\mathbf{x}}(u)=\frac{\sqrt{\kappa_{\mathbf{y}}^{2}+\tau_{\mathbf{y}}^{2}}}{(c-s) \kappa_{\mathbf{y}}} \tag{15}
\end{equation*}
\]

Let \(M\) be a regular surface in \(\mathbb{R}^{3}\) parameterized by \(\chi(u, v)\). Some basic concepts of \(M\) surface is studied by M.P. Do Cormo and these concepts are given below 3.

The standart unit normal vector field \(\mathbf{n}\) on surface \(M\) can be defined by
\[
\begin{equation*}
\mathbf{n}=\frac{\chi_{u} \times \chi_{v}}{\left\|\chi_{u} \times \chi_{v}\right\|} \tag{16}
\end{equation*}
\]

Also, the first and second fundamental forms of the surface \(M\) are as follows
\[
\begin{aligned}
& I=E d u^{2}+2 F d u d v+G d v^{2} \\
& I I=e d u^{2}+2 f d u d v+g d v^{2}
\end{aligned}
\]
where the \(E, F\) and \(G\) components are called the coefficients of the first fundemental form of the surface, and the \(e, f\) and \(g\) components are called the coefficients of the second fundemental form, respectively. The following equations are given for the first and second fundamental form coefficients of the surface
\[
\begin{equation*}
E=\left\langle\chi_{u}, \chi_{u}\right\rangle, \quad F=\left\langle\chi_{u}, \chi_{v}\right\rangle, \quad G=\left\langle\chi_{v}, \chi_{v}\right\rangle \tag{17}
\end{equation*}
\]
and
\[
\begin{equation*}
e=\left\langle\chi_{u u}, \mathbf{n}\right\rangle, \quad f=\left\langle\chi_{u v}, \mathbf{n}\right\rangle, \quad g=\left\langle\chi_{v v}, \mathbf{n}\right\rangle \tag{18}
\end{equation*}
\]

On the other hand, the Gaussian curvature \(K\) and the mean curvature \(H\) of the surface \(M\) are as follows
\[
\begin{equation*}
K=\frac{e g-f^{2}}{E G-F^{2}} \tag{19}
\end{equation*}
\]
and
\[
\begin{equation*}
H=\frac{E g+G e-2 F f}{2\left(E G-F^{2}\right)} \tag{20}
\end{equation*}
\]

Theorem 1. Let \(M\) be a regular surface in \(\mathbb{R}^{3}\). If the Gaussian curvature of the surface \(M\) is zero, the surface is called the developable surface [26] .

Theorem 2. Let \(M\) be a regular surface in \(\mathbb{R}^{3}\). If the mean curvature of the surface \(M\) is zero, the surface is called the minimal surface [26].

\section*{3. Translation Surfaces Created with Curve Pairs}

Translation surfaces are formed by the sum of the two curves, from Eq. (1), translation surface is as follows
\[
\begin{equation*}
\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v) \tag{21}
\end{equation*}
\]
where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curves. If the partial derivatives of the translation surface given above are taken according to \(u\) and \(v\), we have
\[
\begin{align*}
\chi_{u} & =\mathbf{t}_{\mathbf{x}}  \tag{22}\\
\chi_{v} & =\mathbf{t}_{\mathbf{y}}  \tag{23}\\
\chi_{u u} & =\kappa_{\mathbf{x}} \mathbf{n}_{\mathbf{x}}  \tag{24}\\
\chi_{v v} & =\kappa_{\mathbf{y}} \mathbf{n}_{y}  \tag{25}\\
\chi_{u v} & =\frac{d}{d v} \mathbf{t}_{\mathbf{x}} \tag{26}
\end{align*}
\]

The unit normal of the translation surface from Eqs. (16), (22) and (23), we get
\[
\begin{equation*}
\mathbf{n}=\frac{\mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}}{\left\|\mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\|} \tag{27}
\end{equation*}
\]

The coefficients of the first and second fundamental forms of the translation surface are obtained from Eqs. (17), (18) and Eqs. (22)-(26), as
\[
\begin{align*}
E & =\left\langle\chi_{u}, \chi_{u}\right\rangle=1  \tag{28}\\
F & =\left\langle\chi_{u}, \chi_{v}\right\rangle=\left\langle\mathbf{t}_{\mathbf{x}}, \mathbf{t}_{\mathbf{y}}\right\rangle  \tag{29}\\
G & =\left\langle\chi_{v}, \chi_{v}\right\rangle=1 \tag{30}
\end{align*}
\]
and
\[
\begin{align*}
e & =\left\langle\chi_{u u}, \mathbf{n}\right\rangle=\frac{\kappa_{\mathbf{x}}}{\left\|\mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\|}\left\langle\mathbf{n}_{\mathbf{x}}, \mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\rangle  \tag{31}\\
f & =\left\langle\chi_{u v}, \mathbf{n}\right\rangle=\frac{1}{\left\|\mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\|}\left\langle\frac{d}{d v} \mathbf{t}_{\mathbf{x}}, \mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\rangle,  \tag{32}\\
g & =\left\langle\chi_{v v}, \mathbf{n}\right\rangle=\frac{\kappa_{\mathbf{y}}}{\left\|\mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\|}\left\langle\mathbf{n}_{\mathbf{y}}, \mathbf{t}_{\mathbf{x}} \times \mathbf{t}_{\mathbf{y}}\right\rangle . \tag{33}
\end{align*}
\]
3.1. Let \(\mathbf{x}\) and \(\mathbf{y}\) Bertrand partner curves. Let the curves \(\mathbf{x}\) and \(\mathbf{y}\), which are the generating curves of the translation surface parameterized by Eq. (1), be the Bertrand partner curve. In this case, from Eq. (2) and (27) the unit normal of the translation surface is
\[
\begin{equation*}
\mathbf{n}=\frac{\left(\cos \theta \mathbf{t}_{\mathbf{y}}-\sin \theta \mathbf{b}_{\mathbf{y}}\right) \times \mathbf{t}_{\mathbf{y}}}{\left\|\left(\cos \theta \mathbf{t}_{\mathbf{y}}-\sin \theta \mathbf{b}_{\mathbf{y}}\right) \times \mathbf{t}_{\mathbf{y}}\right\|}=-\mathbf{n}_{\mathbf{y}} \tag{34}
\end{equation*}
\]

Since the principal normal vector fields of Bertrand curve pairs are linearly dependent, at the same time \(\mathbf{n}=-\mathbf{n}_{\mathbf{x}}\).

The coefficients of the first fundamental form from Eq. (2) and Eqs. (28)-(30), are obtained as
\[
\begin{aligned}
E & =\left\langle\chi_{u}, \chi_{u}\right\rangle=1 \\
F & =\left\langle\chi_{u}, \chi_{v}\right\rangle=\left\langle\left(\cos \theta \mathbf{t}_{\mathbf{y}}-\sin \theta \mathbf{b}_{\mathbf{y}}\right), \mathbf{t}_{\mathbf{y}}\right\rangle=\cos \theta \\
G & =\left\langle\chi_{v}, \chi_{v}\right\rangle=1
\end{aligned}
\]

The coefficients of the second fundamental form from Eqs. (2), (5) and Eqs. (31)(33), are as follows
\[
\begin{aligned}
e & =\left\langle\kappa_{\mathbf{x}} \mathbf{n}_{\mathbf{x}},-\mathbf{n}_{\mathbf{x}}\right\rangle=-\kappa_{\mathbf{x}} \\
f & =\left\langle\left(\kappa_{\mathbf{y}} \cos \theta+\tau_{\mathbf{y}} \sin \theta\right) \mathbf{n}_{\mathbf{y}},-\mathbf{n}_{\mathbf{y}}\right\rangle=-\kappa_{\mathbf{x}} \\
g & =\left\langle\kappa_{\mathbf{y}} \mathbf{n}_{\mathbf{y}},-\mathbf{n}_{\mathbf{y}}\right\rangle=-\kappa_{\mathbf{y}}
\end{aligned}
\]

The Gaussian and mean curvatures of translation surfaces, whose generating curves are Bertrand partner curves from Eqs. (19) and (20), are calculated as
\[
\begin{equation*}
K=\frac{e g-f^{2}}{E G-F^{2}}=\frac{\kappa_{\mathbf{x}}\left(\kappa_{\mathbf{y}}-\kappa_{\mathbf{x}}\right)}{\sin ^{2} \theta} \tag{35}
\end{equation*}
\]
and
\[
\begin{equation*}
H=\frac{E g+G e-2 F f}{2\left(E G-F^{2}\right)}=\frac{-\kappa_{\mathbf{x}}-\kappa_{\mathbf{y}}+2 \cos \theta \kappa_{\mathbf{x}}}{2 \sin ^{2} \theta} \tag{36}
\end{equation*}
\]

Theorem 3. Let \(\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v)\) be a translation surface where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curve. For translation surfaces, whose generating curves are Bertrand partner curves to be developable surfaces the necessary and sufficient condition is that this \(\mathbf{y}\) is helix.

Proof. Considering that \(\kappa_{\mathbf{x}} \neq 0\), from Eqs. (5), (35) and Theorem 1 , it becomes
\[
\kappa_{\mathbf{x}}=\kappa_{\mathbf{y}}
\]
and
\[
\kappa_{\mathbf{y}} \cos \theta+\tau_{\mathbf{y}} \sin \theta=\kappa_{\mathbf{y}}
\]

So, we get
\[
\frac{\tau_{\mathbf{y}}}{\kappa_{\mathbf{y}}}=\frac{1-\cos \theta}{\sin \theta}
\]

Since \(\theta\) is a constant angle, \(\frac{\tau_{\mathbf{y}}}{\kappa_{\mathbf{y}}}=\) constant. So generating curve \(\mathbf{y}\) is helix.
Theorem 4. Let \(\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v)\) be a translation surface where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curves. Suppose that the generating curves are a pair of Bertrand curves. The necessary and sufficient condition for the surface \(\chi\) to be a minimal surface is that the curve \(\mathbf{x}\) is a helix.

Proof. From Eqs. (5) and (6), we can easily see that
\[
\begin{equation*}
\kappa_{\mathbf{y}}=\kappa_{\mathbf{x}}(v) \cos \theta-\tau_{\mathbf{x}}(v) \sin \theta \tag{37}
\end{equation*}
\]

Using Eqs. (36), (37) and Theorem 2, the following equation can be given
\[
\kappa_{\mathbf{x}}-\cos \theta \kappa_{\mathbf{x}}=\tau_{\mathbf{x}} \sin \theta
\]
and
\[
\frac{\tau_{\mathbf{x}}}{\kappa_{\mathrm{x}}}=\frac{1-\cos \theta}{\sin \theta}
\]

Since \(\theta\) is a constant angle, \(\frac{\tau_{\mathbf{x}}}{\kappa_{\mathbf{x}}}=\) constant. So generating curve \(\mathbf{x}\) is helix.
Example 1. Let \(\mathbf{x}: I \subset \mathbb{R} \rightarrow \mathbb{E}^{3}\) be naturally parametrized curve in \(\mathbb{R}^{3}\) parameterized by
\[
\mathbf{x}(u)=\left(\cos \frac{u}{5}, \sin \frac{u}{5}, \frac{\sqrt{24}}{5} u\right)
\]

The naturally parametrized curve \(\mathbf{y}\) which is the Bertrand partner curve of the \(\mathbf{x}\) curve is as follows
\[
\mathbf{y}(v)=\left(\frac{24}{25} \cos \frac{v}{5}, \frac{24}{25} \sin \frac{v}{5}, \frac{\sqrt{24}}{5} v\right) .
\]

The translation surface generating by the \(\mathbf{x}\) and \(\mathbf{y}\) Bertrand partner curves is parameterized as follows
\[
\chi(u, v)=\left(\cos \frac{u}{5}+\frac{24}{25} \cos \frac{v}{5}, \sin \frac{u}{5}+\frac{24}{25} \sin \frac{v}{5}, \frac{\sqrt{24}}{5} u+\frac{\sqrt{24}}{5} v\right)
\]

In Fig. (1), we present the graph of the above translation surface and its generating Bertrand partner curves \(\mathbf{x}\) and \(\mathbf{y}\).


Figure 1. Translation surface and its generating curves \(\mathbf{x}\) (Red) and \(\mathbf{y}\) (Blue) for Bertrand partner curve.
3.2. Let \(\mathbf{x}\) and \(\mathbf{y}\) Mannheim partner curves. Let the curves \(\mathbf{x}\) and \(\mathbf{y}\), which are the generating curves of the translation surface parameterized by Eq. (1), be the Mannheim partner curves. In this case, from Eq. (7) and (27), the unit normal of the translation surface is
\[
\begin{equation*}
\mathbf{n}=\frac{\left(\cos \theta \mathbf{t}_{\mathbf{y}}+\sin \theta \mathbf{n}_{\mathbf{y}}\right) \times \mathbf{t}_{\mathbf{y}}}{\left\|\left(\cos \theta \mathbf{t}_{\mathbf{y}}+\sin \theta \mathbf{n}_{\mathbf{y}}\right) \times \mathbf{t}_{\mathbf{y}}\right\|}=-\mathbf{b}_{\mathbf{y}} \tag{38}
\end{equation*}
\]

Since the principal normal vector and binormal vector fields of Mannheim curve pairs are linearly dependent, at the same time \(\mathbf{n}=-\mathbf{n}_{\mathbf{x}}\). The coefficients of the first fundamental form from Eq. (7) and Eqs. (28)-(30), are as follow
\[
\begin{aligned}
E & =\left\langle\chi_{u}, \chi_{u}\right\rangle=1 \\
F & =\left\langle\chi_{u}, \chi_{v}\right\rangle=\left\langle\left(\cos \theta \mathbf{t}_{\mathbf{y}}+\sin \theta \mathbf{n}_{\mathbf{y}}\right), \mathbf{t}_{\mathbf{y}}\right\rangle=\cos \theta \\
G & =\left\langle\chi_{v}, \chi_{v}\right\rangle=1
\end{aligned}
\]

The coefficients of the second fundamental form from Eqs. (7), (10) and Eqs. (31)(33), are obtained as
\[
\begin{aligned}
e & =\left\langle\kappa_{\mathbf{x}} \mathbf{n}_{\mathbf{x}},-\mathbf{n}_{\mathbf{x}}\right\rangle=-\kappa_{\mathbf{x}} \\
f & =\left\langle-\kappa_{\mathbf{y}} \sin \theta \mathbf{t}_{\mathbf{y}}+\kappa_{\mathbf{y}} \cos \theta \mathbf{n}_{\mathbf{y}}+\tau_{\mathbf{y}} \sin \theta \mathbf{b}_{\mathbf{y}},-\mathbf{b}_{\mathbf{y}}\right\rangle=-\tau_{\mathbf{y}} \sin \theta \\
g & =\left\langle\kappa_{\mathbf{y}} \mathbf{n}_{\mathbf{y}}, \mathbf{b}_{\mathbf{y}}\right\rangle=0
\end{aligned}
\]

If we calculate the Gaussian and mean curvatures of translation surfaces, whose generating curves are Mannheim partner curves, from Eqs. (19) and 20, we have
\[
\begin{equation*}
K=\frac{e g-f^{2}}{E G-F^{2}}=-\tau_{\mathbf{y}}^{2} \tag{39}
\end{equation*}
\]
and
\[
\begin{equation*}
H=\frac{E g+G e-2 F f}{2\left(E G-F^{2}\right)}=\frac{-\kappa_{\mathbf{x}}+\tau_{\mathbf{y}} \sin 2 \theta}{2 \sin ^{2} \theta} \tag{40}
\end{equation*}
\]

Theorem 5. Let \(\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v)\) be a translation surface where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curve. For translation surfaces, whose generating curves are Mannheim partner curves to be developable surfaces, the necessary sufficient condition is that the curve \(\mathbf{y}\) is a planar curve.
Proof. It is easily seen from Eq. (39) and Theorem 1 that \(\tau_{\mathbf{y}}=0\). This means that the curve \(\mathbf{y}\) is a planar curve.

Theorem 6. Let \(\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v)\) be a translation surface where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curve. For translation surfaces, whose generating curves are Mannheim partner curves to be minimal surfaces, the necessary sufficient condition is that the curve \(\mathbf{y}\) is a planar curve or \(v=c_{1} u+c_{2}, c_{1}, c_{2} \in \mathbb{R}\).

Proof. From Eqs. (10), 40) and Theorem 2, the following equation can be given
\[
\tau_{\mathbf{y}} \sin 2 \theta=\tau_{\mathbf{y}} \sin \theta \frac{d v}{d u}
\]
and
\[
2 \tau_{\mathbf{y}} \cos \theta=\tau_{\mathbf{y}} \frac{d v}{d u}
\]

Here \(\tau_{\mathbf{y}}=0\) is an obvious solution. So \(\mathbf{y}\) is a planar curve. Let \(\tau_{\mathbf{y}} \neq 0\) then, we get
\[
2 \cos \theta \int d u=\int d v
\]

If \(2 \cos \theta=c_{1}, c_{1} \in \mathbb{R}\) is selected here, we obtain
\[
v=c_{1} u+c_{2}, \quad c_{1}, c_{2} \in \mathbb{R}
\]

Example 2. Let \(\mathbf{x}: I \subset \mathbb{R} \rightarrow \mathbb{E}^{3}\) be arbitrary parametrized curve in \(\mathbb{R}^{3}\) parameterized by
\[
\mathbf{x}(u)=\left(\frac{8}{5} \cos u, \frac{8}{5} \sin u, \frac{4}{5} u\right)
\]

The arbitrary parametrized curve \(\mathbf{y}\) which is the Mannheim partner curve of the curve \(\mathbf{x}\) is as follows
\[
\mathbf{y}(v)=\left(-\frac{8}{5}(\sin v+\cos v), \frac{8}{5}(\sin v+\cos v), \frac{4}{5} v\right) .
\]

The translation surface generating by the \(\mathbf{x}\) and \(\mathbf{y}\) Mannheim partner curves is parameterized as follows
\[
\chi(u, v)=\left(\frac{8}{5} \cos u-\frac{8}{5}(\sin v+\cos v), \frac{8}{5} \sin u+\frac{8}{5}(\sin v+\cos v), \frac{4}{5} u+\frac{4}{5} v\right) .
\]

In Fig. (2), we present the graph of the above translation surface and its generating Mannheim partner curves \(\mathbf{x}\) and \(\mathbf{y}\).
3.3. Let \(\mathbf{x}\) and \(\mathbf{y}\) involute-evolute partner curves. Let the curves \(\mathbf{x}\) and \(\mathbf{y}\), which are the generating curves of the translation surface parameterized by Eq. (1), be the involute-evolute partner curves. So, from Eq. (12) and (27), the unit normal of the translation surface is
\[
\begin{equation*}
\mathbf{n}=\frac{\mathbf{n}_{\mathbf{y}} \times \mathbf{t}_{\mathbf{y}}}{\left\|\mathbf{n}_{\mathbf{y}} \times \mathbf{t}_{\mathbf{y}}\right\|}=-\mathbf{b}_{\mathbf{y}} \tag{41}
\end{equation*}
\]

The coefficients of the first fundamental form from Eq. (12) and Eqs. (28)-(30), are as follows
\[
\begin{aligned}
E & =\left\langle\chi_{u}, \chi_{u}\right\rangle=1 \\
F & =\left\langle\chi_{u}, \chi_{v}\right\rangle=\left\langle\mathbf{n}_{\mathbf{y}}, \mathbf{t}_{\mathbf{y}}\right\rangle=0 \\
G & =\left\langle\chi_{v}, \chi_{v}\right\rangle=1
\end{aligned}
\]

If we calculate the coefficients of the second fundamental form from Eqs. (12), 13), (15) and Eqs. (31)-(33), we can easily see that
\[
e=\left\langle\kappa_{\mathbf{x}} \mathbf{n}_{\mathbf{x}},-\mathbf{b}_{\mathbf{y}}\right\rangle=-\kappa_{\mathbf{x}} \sin \theta
\]


Figure 2. Translation surface and its generating curves \(\mathbf{x}(\) Red \()\) and \(\mathbf{y}\) (Blue) for Mannhiem partner curve.
\[
\begin{aligned}
f & =\left\langle-\kappa_{\mathbf{y}} \mathbf{t}_{\mathbf{y}}+\tau_{\mathbf{y}} \mathbf{b}_{\mathbf{y}},-\mathbf{b}_{\mathbf{y}}\right\rangle=-\tau_{\mathbf{y}} \\
g & =\left\langle\kappa_{\mathbf{y}} \mathbf{n}_{\mathbf{y}}, \mathbf{b}_{\mathbf{y}}\right\rangle=0
\end{aligned}
\]

The Gaussian and mean curvatures of translation surfaces, whose generating curves are involute-evolute partner curves are obtained from Eqs. (19) and (20), as follows
\[
\begin{equation*}
K=\frac{e g-f^{2}}{E G-F^{2}}=-\tau_{\mathbf{y}}^{2} \tag{42}
\end{equation*}
\]
and
\[
\begin{equation*}
H=\frac{E g+G e-2 F f}{2\left(E G-F^{2}\right)}=\frac{-\kappa_{\mathbf{x}} \sin \theta}{2} \tag{43}
\end{equation*}
\]

Theorem 7. Let \(\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v)\) be translation surface where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curves. Suppose that the generating curves are the involute-evolute partner curves. The necessary and sufficient condition for the surface \(\chi\) to be developable surface is that the curve \(\mathbf{y}\) is a planar curve.

Proof. It is easily seen from Eq. (42) and Theorem 1 that \(\tau_{\mathbf{y}}=0\). This means that the curve \(\mathbf{y}\) is a planar curve.

Theorem 8. Let \(\chi(u, v)=\mathbf{x}(u)+\mathbf{y}(v)\) be translation surface where \(\mathbf{x}\) and \(\mathbf{y}\) are generating curves. Suppose that the generating curves are the involute-evolute partner curves. In this case, the translation surface \(\chi\) cannot be a minimal surface.

Proof. Since \(\kappa_{\mathbf{x}} \neq 0\), considering Eq. (43), it is seen that \(H \neq 0\). Therefore, such translation surfaces cannot be minimal.
Example 3. Let \(\mathbf{x}: I \subset \mathbb{R} \rightarrow \mathbb{E}^{3}\) be arbitrary parametrized curve in \(\mathbb{R}^{3}\) parameterized by
\[
\mathbf{x}(u)=\left(\frac{8}{5} \cos u, \frac{8}{5} \sin u, \frac{4}{5} u\right)
\]

The arbitrary parametrized curve \(\mathbf{y}\) involute partner curve of the \(\mathbf{x}\) curve is as follows
\[
\mathbf{y}(v)=\left(\frac{8}{5} \cos v-\frac{2}{5} \sin v+\frac{2}{5} v \sin v, \frac{8}{5} \sin v+\frac{2}{5} \cos v-\frac{2}{5} v \cos v, \frac{3}{5} v\right) .
\]

The translation surface generating by the \(\mathbf{x}\) and \(\mathbf{y}\) involute-evolute partner curves is parameterized as follows
\(\chi(u, v)=\left(\frac{8}{5} \cos u+\frac{8}{5} \cos v-\frac{2}{5} \sin v+\frac{2}{5} v \sin v, \frac{8}{5} \sin u+\frac{8}{5} \sin v+\frac{2}{5} \cos v 2 \frac{2}{5} v \cos v, \frac{4}{5} u+\frac{3}{5} v\right)\).
In Fig. (3), we present the graph of the above translation surface and its generating involute-evolute partner curves \(\mathbf{x}\) and \(\mathbf{y}\).


Figure 3. Translation surface and its generating curves \(\mathbf{x}\) (Red) and \(\mathbf{y}\) (Blue) for involute-evolute partner curves.

Author Contribution Statements The authors jointly worked on the results and they read and approved the final manuscript.

Declaration of Competing Interests The authors declare that they have no competing interest.

Acknowledgements The authors thank the referees for their valuable contributions to the article.

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FRACTIONAL ORDER MATHEMATICAL MODELING OF LUMPY SKIN DISEASE


M otivation and Research B ackground.


Structure of the Paper.
\(\square\)


Def nition \(1 \square\). The Caputo fractional derivative of a continuous function \(g\) on,\(T\) is defined as:
\[
\mathfrak{D}^{\alpha} g t \quad \frac{}{n-\alpha} \int^{t} t-s^{n-\alpha-} \frac{d^{n}}{d s^{n}} g s d s
\]
where \(<\alpha \leq, n \quad \alpha \quad\), and \(\alpha\) represents the integer part of \(\alpha\).

Def nition \(2 \square\). The fractional integral of a continuous function \(g\) on \(L \quad, T, \mathbb{R}\) of order \(<\alpha \leq\) corresponding to \(t\) is defined as:
\[
I^{\alpha} g t \quad-\quad \int^{t} t-s^{\alpha-} g s d s
\]

Def nition \(3 \square\). The Laplace transform is defined by
\[
F s \quad L f t \quad \int^{\infty} e^{-s t} f t d t
\]
where \(f t\) is \(n\)-dimensional vector-valued function.
Def nition \(4 \square\). The Mittag-Leffler function in two parameters is defined as
\[
E_{\alpha, \beta} z \quad \sum_{k}^{\infty} \frac{z^{k}}{\alpha k \quad \beta}, z \in \mathbb{C}
\]
where \(\alpha>, \beta>, \mathbb{C}\) denotes the complex plane.
Lemma \(1 \square\). Let \(\mathbb{C}\) be a complex plane, for any \(\alpha>, \beta>\) and \(A \in C^{n \times n}\),
\[
L t^{\beta-} E_{\alpha, \beta} A t^{\alpha} \quad \frac{s^{\alpha-\beta}}{s^{\alpha}-A}
\]
holds for Re \(s>\|A\|^{\frac{1}{\alpha}}\), where Re s represents the real part of the complex number \(s\).

Lemma \(2 \square\). Let \(F s\) be the Laplace transform of the function \(f t, n\) being an integer then the Laplace transform of the Caputo fractional derivative of order \(\alpha\) is given by
\[
L \mathfrak{D}^{\alpha} f t \quad s^{\alpha} F s-\sum_{k}^{n} s^{\alpha-k} f^{k-} \quad, n-\quad<\alpha \leq n .
\]

Lemma 3
. Let \(g t \in C a, b\) and \(\mathfrak{D}^{\alpha} g t \in\)
\(C a, b\) for \(<\alpha \leq\), then
\[
g t \quad g a \quad-\quad \mathfrak{D}^{\alpha} g \quad s \quad t-a^{\alpha}
\]
with \(\leq s \leq t, \forall t \in a, b\). Thus, we can deduce that for \(g t \in C, b\) and Caputo fractional derivative \(\mathfrak{D}^{\alpha} g t \in C, b\) for \(<\alpha \leq\), if \(\mathfrak{D}^{\alpha} g t \geq, \forall t \in, b\), then the function \(g t\) is non-decreasing and if \(\mathfrak{D}^{\alpha} g t \leq, \forall t \in, b\), then the function \(g t\) is non-increasing \(\forall t \in \quad, b\).

Theorem \(1 \square\). Consider the fractional differential equation:
\[
\begin{array}{ll}
\mathfrak{D}^{\alpha} \boldsymbol{x} t & f t, \boldsymbol{x} t, \\
\boldsymbol{x}^{k} t & \boldsymbol{x}^{k}, k \quad,, \ldots, n-,
\end{array}
\]
where \(\mathfrak{D}^{\alpha}\) represents the Caputo fractional derivative. Let \(L>\) and \(f \quad, L \times\) \(\mathbb{R} \rightarrow \mathbb{R}\) is continuous and suppose that there exists a real number \(l>\) such that \(|f t, x-f t, y| \leq l|x-y|\) for \(t \in \quad, L\) and \(x, y \in \mathbb{R}\). Then, the initial value problem has a unique solution in \(A C, L\).
Theorem \(2 \square\). Consider the following fractional-order system:
\[
\mathfrak{D}^{\alpha} X t \quad \mathcal{F} X
\]
with \(<\alpha<, X t \quad x t, x t, \ldots, x^{n} t\) and \(\mathcal{F} X \quad t, \infty \rightarrow \mathbb{R}^{n \times n}\). The equilibrium points of system \(\square\) are evaluated by solving system of equations \(\mathcal{F} X\). These equilibrium points are locally asymptotically stable if each eigenvalue \(\lambda\) of the Jacobian matrix \(J X\) calculated at the equilibrium points satisfies \(\left|\arg \lambda_{i}\right|>\underline{\alpha \pi}\).
\[
\mathcal{N} \quad \mathcal{S} \in \mathcal{E} \quad \begin{array}{llll}
\mathcal{R} \\
& & & \mathcal{S}
\end{array}
\]
\(\mathcal{E}\)
\(\mathcal{I}\)
\(\mathcal{R}\)
\(\beta \quad \eta \rho \sigma\)
\(\mathfrak{D}_{t}^{\alpha} \mathcal{S}_{t} \quad-\beta \mathcal{S}_{t} \mathcal{I}_{t}-\sigma \mathcal{S}_{t}\),
\(\mathfrak{D}_{t}^{\alpha} \mathcal{E}_{t} \quad \beta \mathcal{S}_{t} \mathcal{I}_{t}-\sigma \quad \eta \mathcal{E}_{t}\),
\(\mathfrak{D}_{t}^{\alpha} \mathcal{I}_{t} \quad \eta \mathcal{E}_{t}-\rho \quad \sigma \mathcal{I}_{t}\),
\(\mathfrak{D}_{t}^{\alpha} \mathcal{R}_{t} \quad \rho \mathcal{I}_{t}-\sigma \mathcal{R}_{t}\)
\(\begin{array}{llllllllll}\mathcal{S}_{t} & \mathcal{S} & \mathcal{E}_{t} & \mathcal{E} & \mathcal{I}_{t} & \mathcal{I} & \mathcal{R}_{t} & & \mathcal{R} \\ & & & \alpha & . & <\alpha< & & \end{array}\)
\(\mathfrak{D}_{t}^{\alpha}\)
\[
\begin{gathered}
\beta \\
\eta \\
\rho \\
\sigma \\
\hline
\end{gathered}
\]


Theorem 3. There is a unique solution \(\mathcal{U} t \quad \mathcal{S} t, \mathcal{E} t, \mathcal{I} t, \mathcal{R} t^{T}\) for the initial value problem given by the system of equations in \(\square\) on \(t \geq\) in , \(\theta\) and the solution will remain in \(\mathbb{R}\). Furthermore, the solutions are all bounded.

Proof.
\[
\begin{array}{cc}
, \infty \\
\left.\mathfrak{D}_{t}^{\alpha} \mathcal{S}_{t}\right|_{\mathcal{S}} & > \\
\left.\mathfrak{D}_{t}^{\alpha} \mathcal{E}_{t}\right|_{\mathcal{E}} & \beta \mathcal{S}_{t} \mathcal{I}_{t} \geq \\
\mathfrak{D}_{t}^{\alpha} \mathcal{I}_{t} \mathcal{I} & \eta \mathcal{E}_{t} \geq, \\
\left.\mathfrak{D}_{t}^{\alpha} \mathcal{R}_{t}\right|_{\mathcal{R}} & \rho \mathcal{I}_{t} \geq
\end{array}
\]
\(\mathbb{R}\)
\[
\mathfrak{D}^{\alpha} \mathcal{N} t \quad-\sigma \mathcal{N} t \geq
\]
\[
\text { i.e. } \mathfrak{D}^{\alpha} \mathcal{N} t \quad \sigma \mathcal{N} t \leq
\]
\[
\begin{aligned}
& \text { ㅁ } \\
& \left\{\mathcal{S}_{t}, \mathcal{E}_{t}, \mathcal{I}_{t}, \mathcal{R}_{t} \in \mathbb{R} \quad \mathcal{S}, \mathcal{E}, \mathcal{I}, \mathcal{R} \geq \quad \leq \mathcal{S}_{t} \quad \mathcal{E}_{t} \quad \mathcal{I}_{t} \quad \mathcal{R}_{t} \leq \frac{-}{\sigma}\right\} . \\
& \mathcal{N} t \quad \mathcal{S} t \quad \mathcal{E} t \quad \mathcal{I} t \mathcal{R} t \\
& \alpha \\
& \mathfrak{D}^{\alpha} \mathcal{N} t \quad \mathfrak{D}^{\alpha} \mathcal{S}_{t} \quad \mathfrak{D}^{\alpha} \mathcal{E}_{t} \quad \mathfrak{D}^{\alpha} \mathcal{I}_{t} \quad \mathfrak{D}^{\alpha} \mathcal{R}_{t} \\
& \mathfrak{D}^{\alpha} \mathcal{N} t \quad-\sigma \mathcal{N} t . \\
& \mathcal{N} s \quad \frac{s^{-} s^{\alpha-} \mathcal{N}}{s^{\alpha} \sigma} . \\
& \square \\
& \mathcal{N} t \quad \bar{\sigma} \quad-E_{\alpha}-\sigma t^{\alpha} \quad \mathcal{N} \quad E_{\alpha}-\sigma t^{\alpha} . \\
& E_{\alpha}-t \quad t>\quad \leq E_{\alpha}-\sigma t^{\alpha} \leq \\
& <\alpha \leq \square \square \\
& N t \leq \frac{\bar{\sigma}}{} . \\
& t \geq \quad \mathcal{S} t \mathcal{E} t \mathcal{I} t \mathcal{R} t \\
& \text {, } \infty
\end{aligned}
\]

Equilibrium Points.
LSD-free equilibrium.
\[
\begin{aligned}
& \mathcal{E} \\
&(\bar{\sigma},,,) \mathcal{I}^{\mathcal{I}_{t}} \\
& \mathcal{R}
\end{aligned}
\]

Reproduction Number:

\(\mathcal{V}\)
\(\mathcal{V} t \mathcal{V}^{-} t-\mathcal{V} t\)
\[
\mathcal{F} t\left[\begin{array}{lll}
\beta \mathcal{S}_{t} \mathcal{I}_{t}
\end{array}\right], \mathcal{V} \quad t \quad\left[\begin{array}{l} 
\\
\mathcal{F} \mathcal{E}_{t} \\
\rho \mathcal{I}_{t}
\end{array}\right], \mathcal{V}^{-} t \quad\left[\begin{array}{ccc}
\beta \mathcal{S}_{t} \mathcal{I}_{t} & \sigma \mathcal{S}_{t} \\
\sigma & \eta & \mathcal{E}_{t} \\
\rho & \sigma & \mathcal{I}_{t} \\
\sigma \mathcal{R}_{t}
\end{array}\right] .
\]

E
\[
\mathcal{F}^{*} t\left[\begin{array}{ll}
\frac{\beta}{\sigma} \\
&
\end{array}\right]
\]
\[
\begin{gathered}
\mathcal{V} t \\
\mathcal{V}^{*} t \quad \mathcal{V}^{*-} t-\mathcal{V}^{*} t
\end{gathered}\left[\begin{array}{ccccc}
\sigma & & & \frac{\beta}{\sigma} & \\
& \sigma & \eta & & \\
& -\eta & \rho & \sigma & \\
& & & -\rho & \sigma
\end{array}\right] .
\]
\(\mathcal{F}^{*} \mathcal{V}^{*-}\)
\[
\begin{aligned}
& \mathcal{F}^{*} \mathcal{V}^{*-} \\
& R \quad \frac{\beta \quad \eta}{\sigma \sigma \quad \eta \quad \sigma \quad \rho} .
\end{aligned}
\]

A nalyzing \(R\) :
\[
R
\]
\[
\begin{aligned}
& \text { R } \\
& \frac{\partial R}{\partial \beta} \quad \begin{array}{c}
\text { 白 } \sigma \quad \eta \quad \sigma \quad \rho
\end{array}, \\
& \frac{\partial R}{\partial} \quad \frac{\beta \eta}{\sigma \sigma \quad \eta \sigma \quad \rho}>, \\
& \frac{\partial R}{\partial \eta} \quad \begin{array}{lll}
\sigma \sigma & \beta & \sigma \\
\sigma \sigma & \sigma
\end{array}, \\
& \frac{\partial R}{\partial \rho} \quad \frac{-\beta}{\sigma \sigma} \quad \eta \quad \sigma \quad \rho \quad<, \\
& \frac{\partial R}{\partial \sigma} \quad \frac{-\beta}{\sigma} \quad \begin{array}{lll} 
& \eta & \sigma
\end{array} \quad \rho\left\{\begin{array}{lll}
\sigma & & \\
\sigma & \eta & \sigma \quad \rho
\end{array}\right\}<. \\
& \beta \quad \eta \\
& \rho \quad \sigma
\end{aligned}
\]
\[
\begin{aligned}
& \text { I / } \\
& \mathcal{S}, \mathcal{E}, \mathcal{I}, \mathcal{R} \\
& \mathcal{S} \quad \frac{\sigma \quad \eta \quad \rho \quad \rho}{\beta \eta}, \mathcal{E} \quad \frac{\beta \quad \eta-\sigma \sigma}{} \eta \quad \begin{array}{l}
\beta \eta \\
\beta \eta \\
\hline
\end{array}, \\
& \mathcal{I} \quad \frac{\sigma R-}{\beta} \quad \mathcal{R} \quad \frac{\rho R-}{\beta} \\
& \mathcal{S}, \mathcal{E}, \mathcal{I}, \mathcal{R} \quad>\quad R>
\end{aligned}
\]

\section*{Stability A nalysis.}

Theorem 4. LSD-free equilibrium point \(E \quad(\bar{\sigma}\), , , \()\) of the system is locally asymptotically stable when \(R<\), unstable otherwise.
Proof.
E
\[
\left[\begin{array}{cccc}
-\sigma & & & -\frac{\beta}{\sigma} \\
& -\sigma & \eta & \frac{\beta^{\frac{\beta}{\sigma}}}{} \\
& \eta & -\rho^{\sigma} & \\
& & \rho & -\sigma
\end{array}\right]
\]
\[
P \lambda \quad \lambda \quad P \lambda \quad P
\]


Theorem 5. The LSD-persistent equilibrium point \(E \quad \mathcal{S}, \mathcal{E}, \mathcal{I}, \mathcal{R}\) exists and is locally asymptotically stable iff \(R>\).
Proof.
E
\[
\left[\begin{array}{cccc}
-\sigma R & & -\frac{\eta \sigma \rho \sigma}{\eta \eta^{\eta} \sigma} & \\
\sigma R- & -\sigma \quad \eta & \frac{\eta \sigma{ }^{\eta} \sigma}{2} & \\
& \eta & -\rho^{\rho \sigma} & -\sigma
\end{array}\right]
\]
\[
\square
\]
■

\[
\begin{aligned}
& P \lambda \quad \lambda \quad P \lambda \quad P \lambda \quad P \text {, } \\
& P \quad \sigma R \quad \eta \quad \rho \text {, } \\
& P \quad \sigma R \quad \sigma \quad \eta \quad \rho \text {, } \\
& P \quad R-\sigma \eta \sigma \rho \sigma \\
& P>\quad P>\quad R>\quad P P-P> \\
& E \quad R> \\
& R>
\end{aligned}
\]
\begin{tabular}{llllll}
\hline Parameter & Value & & \multicolumn{1}{c}{ Source } \\
\hline & & & & \\
\(\beta\) & & & & \\
\(\eta\) & & & & \\
\(\rho\) & & & & \\
\(\sigma\) & & & & \\
\hline & & & & & \\
\hline Population & \(\mathcal{S}\) & \(\mathcal{E}\) & \(\mathcal{I}\) & \(\mathcal{R}\) \\
\hline Initial Values & & & & \\
\hline
\end{tabular}








\(\alpha\)


\(R\) decreases with increase in death rate \((\sigma)\)
\(R \quad \beta \quad \sigma\)

\(R\) increases with increase in incubation rate ( \(\eta\) )

\(R\) decreases with increase in recovery rate ( \(\rho\) )
\(R \quad \eta \quad \rho\)
\[
R<
\]

A uthor Contribution Statements

Declaration of Competing Interests
http://communications.science.ankara.edu.tr

\title{
SEMIREGULAR, SEMIPERFECT AND SEMIPOTENT MATRIX RINGS RELATIVE TO AN IDEAL
}

\author{
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}

\begin{abstract}
This paper investigates relative ring theoretical properties in the context of formal triangular matrix rings. The first aim is to study the semiregularity of formal triangular matrix rings relative to an ideal. We prove that the formal triangular matrix ring \(T\) is \(T^{\prime}\)-semiregular if and only if \(A\) is \(I\) semiregular, \(B\) is \(K\)-semiregular and \(N=M\) for an ideal \(T^{\prime}=\left(\begin{array}{cc}I & 0 \\ N & K\end{array}\right)\) of \(T=\left(\begin{array}{cc}A & 0 \\ M & B\end{array}\right)\). We also discuss the relative semiperfect formal triangular matrix rings in relation to the strong lifting property of ideals. Moreover, we have considered the behavior of relative semipotent and potent property of formal triangular matrix rings. Several examples are provided throughout the paper in order to highlight our results.
\end{abstract}

\section*{1. Introduction}

The celebrated work of Wedderburn and Artin gave a key insight into the structure of a semisimple artinian ring, which makes it an attractive structure to study. Moreover, for a right artinian ring \(R\), the Jacobson radical is nilpotent, and the ring \(R / \mathrm{J}(R)\) is semisimple, so a main problem would be to "lift" the structure of the factor ring \(R / \mathrm{J}(R)\) onto the ring \(R\) itself. As a consequence of this, we are led to the concept of lifting idempotents and, consequently, to the notion of a semiperfect ring.

Let \(I\) be an ideal in a ring \(R\). Recall that an element \(a \in R\) is an idempotent modulo \(I\) if \(a+I \in R / I\) is an idempotent. In this case, we say that \(a\) can be lifted to an idempotent (modulo \(I\) ) if there exists an idempotent \(e \in R\) with \(e-a \in I\). Note that the ideal \(I\) in \(R\) is called idempotent lifting if, whenever \(a+I \in R / I\) is an idempotent, then there exists an idempotent \(e \in R\) such that \(e-a \in I\).

2020 Mathematics Subject Classification. 16S50; 16D25; 16E50; 16L30.
Keywords. Formal triangular matrix ring; strongly lifting ideal; semiregular ring; semiperfect ring; semipotent ring.
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Lifting of idempotents is a key method in transferring some structural properties of a factor ring of a ring \(R\) up to the ring itself. Several classes of rings are described in terms of the idempotent lifting property of ideals. For example, semiperfect rings are those rings \(R\) for which \(R / \mathrm{J}(R)\) is semisimple and the Jacobson radical \(\mathrm{J}(R)\) of \(R\) is idempotent lifting. Some nontrivial generalizations of semiperfect rings, such as semiregular rings and potent rings may be considered as further examples.

As it has been pointed out above, the idempotent lifting property of the Jacobson radical \(J(R)\) of \(R\) is prominent in the study of semiregular and semiperfect rings. A stronger property than the idempotent lifting property, namely, strong lifting property of ideals, gives rise to a natural generalization of semiregular and semiperfect rings. Semiregular rings relative to an ideal first emerged in a paper 8 by Nicholson and Yousif. Then, Yousif and Zhou 10 studied further semiperfect and perfect rings relative to an ideal in connection with relative semiregular rings. Later, Nicholson and Zhou worked on a natural extension of this work together with strongly lifting ideals to characterize \(I\)-semiregular and \(I\)-semiperfect rings for an ideal \(I\) of a ring \(R\) in 9. Recall that an ideal \(I\) of a ring \(R\) is called strongly lifting if, whenever \(a+I \in R / I\) is an idempotent, then there exists an idempotent \(e \in a R\) such that \(a-e \in I\). In this work, Nicholson and Zhou further showed that the ring \(R\) is \(I\)-semiregular (semiperfect) if and only if \(R / I\) is regular (semisimple) and \(I\) is strongly lifting.

Recall that a ring \(R\) is called semipotent if each one-sided ideal of \(R\) that is not contained in its Jacobson radical \(\mathrm{J}(R)\) contains a nonzero idempotent. A semipotent ring \(R\) is called potent if, in addition, \(\mathrm{J}(R)\) is an idempotent lifting ideal of \(R\). Semipotent rings has been generalized to semipotent rings relative to an ideal by Nicholson and Zhou in 9. It is also important to consider the strong lifting properties of ideals in this setting, and relative potent rings are defined in relation to these ideals as well as relative semipotent rings.

One of important constructions in ring theory is the triangular ring construction. Let \(A, B\) be rings and \(M\) be a \(B-A\) bimodule. A formal triangular matrix ring is a ring of the form
\[
\left(\begin{array}{cc}
A & 0 \\
M & B
\end{array}\right)=\left\{\left.\left(\begin{array}{cc}
a & 0 \\
m & b
\end{array}\right) \right\rvert\, a \in A, b \in B, \text { and } m \in M\right\}
\]
under the usual matrix operations. There are a number of important examples in this class, including lower (upper) triangular matrices over a known ring \(R\). Moreover, many surprising examples and counterexamples have emerged via the triangular ring construction in literature by varying the choices of \(A, B\) and \(M\). By using formal triangular matrix rings, Herstein in 5 provided a counterexample to the Jacobson conjecture, one of the oldest and most well-known conjectures in noncommutative ring theory. In 3 , these rings were studied in detail, and in 4 , various ring theoretic properties of formal triangular matrix rings were investigated.

This paper aims to unify all these relative properties in the framework of formal triangular matrix rings. In Section 2, we completely give a description of the
semiregularity and semiperfectness of formal triangular matrix rings relative to an ideal, proving that \(T\) is \(T^{\prime}\)-semiregular (resp. semiperfect) if and only if \(A\) is \(I\)-semiregular (resp. semiperfect), \(B\) is \(K\)-semiregular (resp. semiperfect) and \(N=M\) for an ideal \(T^{\prime}=\left(\begin{array}{cc}I & 0 \\ N & K\end{array}\right)\) of \(T=\left(\begin{array}{cc}A & 0 \\ M & B\end{array}\right)\) (Theorem 1 and Theorem 2 . Then, we highlight our results by providing several examples, in particular we show that the "if" part of the above theorems are not in general true if we omit the condition \(N=M\). We have further considered the behavior of relative semipotent property of formal triangular matrix rings. Since being a semipotent or a potent ring passes over to formal triangular matrix rings by a result due to Haghany and Varadarajan (4, it is natural to suspect that it may also pass over in the relative case.

Throughout this paper, all rings will be associative rings with an identity element \(1 \neq 0\), not necessarily commutative. We will denote by \(\mathrm{J}(R)\) the Jacobson radical of a ring \(R\).

\section*{2. Results}

Recall that a formal triangular matrix ring \(T\) is a ring of the form
\[
T=\left\{\left.\left(\begin{array}{cc}
a & 0 \\
m & b
\end{array}\right) \right\rvert\, a \in A, b \in B, m \in M .\right\}
\]
under the usual matrix addition and multiplication where \(A, B\) are two rings and \(M\) is a left \(B\) right \(A\) bimodule. For simplicity, we write
\[
T=\left(\begin{array}{cc}
A & 0 \\
M & B
\end{array}\right)
\]
for the formal triangular matrix ring. The construction of examples and counterexamples for asymmetric ring-theoretic properties is among the major applications of such rings in noncommutative ring theory. In particular, 4 provides a comprehensive resource for various ring-theoretic properties of formal triangular matrix rings.

Moreover, Goodearl covered formal triangular matrix rings in his classic book "Ring Theory: Nonsingular rings and modules" 3. In order to better understand the ideal structure of a ring of such a type, we must first recall the following fact.
Proposition 1. 3 If \(I\) is a two-sided ideal of \(A, K\) a two-sided ideal of \(B\), and \(N\) a \(B\) - \(A\) subbimodule of \(M\) which contains \(M I+K M\), then \(\left(\begin{array}{cc}I & 0 \\ N\end{array}\right)\) is a two-sided ideal of \(T\). Conversely, every two-sided ideal of \(T\) has this form.

We will begin by simplifying the following notation: \(T\) denotes the formal triangular matrix ring \(\left(\begin{array}{cc}A & 0 \\ M & B\end{array}\right)\), while \(T^{\prime}\) refers to an ideal of the form \(\left(\begin{array}{cc}I & 0 \\ N\end{array}\right)\) with the additional properties outlined above.

Now we continue with a lemma that is implicit in 8 and proved in 10 Lemma 1.1] and leads us to a number of significant ring-theoretic properties relative to an ideal of a ring \(R\).

Lemma 1. 10, Lemma 1.1] Let I be an ideal of the ring \(R\). The following conditions are equivalent for a right ideal \(I^{\prime}\) of \(R\) :
(1) There exists \(e^{2}=e \in I^{\prime}\) with \((1-e) I^{\prime} \subseteq I\).
(2) There exists \(e^{2}=e \in I^{\prime}\) with \(I^{\prime} \cap(1-e) R \subseteq I\).
(3) \(I^{\prime}=e R \oplus S\) where \(e^{2}=e\) and \(S \subseteq I\).

According to Nicholson and Zhou 9 an ideal \(I\) of the ring \(R\) respects a right ideal \(I^{\prime}\) of \(R\) if the conditions in Lemma 1 are satisfied. Similarly, \(I\) respects a left ideal \(L \subseteq R\) if \(L=R e \oplus S\) where \(e^{2}=e\) and \(S \subseteq I\). It is worth noting that this definition is left-right symmetric, i.e., if \(I \triangleleft R\) and \(a \in R\), then \(I\) respects \(a R\) if and only if \(I\) respects \(R a\).

Right (left) \(I\)-semiregular elements and rings first emerged in a paper 8 by Nicholson and Yousif and then were studied in 10 by Yousif and Zhou. Later, Nicholson and Zhou 9 dealt with these elements in terms of respecting a right (left) ideal as defined above and demonstrated that it is not necessary to distinguish between "right \(I\)-semiregular" and "left \(I\)-semiregular". Let \(I\) be an ideal of the ring \(R\). Recall that an element \(a \in R\) is called \(I\)-semiregular if \(I\) respects \(a R\), i.e., if \(e^{2}=e \in a R\) exists with \((1-e) a \in I\), or alternatively, if \(f^{2}=f \in R a\) exists with \(a(1-f) \in I\). As expected, when all elements of the ring \(R\) are \(I\)-semiregular, the ring \(R\) is called a \(I\)-semiregular ring.

It is well known that the topic of lifting of idempotents is a crucial method for identifying the structure of semiregular and semiperfect rings. Nicholson and Zhou studied a natural extension of these notions in connection with strongly lifting ideals in 9. Recall that an ideal \(I\) of a ring \(R\) is called strongly lifting if, for some \(a \in R\), whenever \(a^{2}-a \in I\), then there exists an idempotent \(e \in a R\) with \(a-e \in I\). It is possible to replace the conclusion \(e \in a R\) by \(e \in R a\) or \(e \in a R a\) since this notion is left-right symmetric 9, Lemma 1]. In this work, Nicholson and Zhou further showed that the ring \(R\) is \(I\)-semiregular if and only if \(R / I\) is regular and \(I\) is strongly lifting. A recent work 1 has shed new light on the question: "What can be said about relative semiregular ideals of the the formal triangular matrix ring?". The author has provided a criterion to decide if a given ideal \(T^{\prime}\) of the formal triangular matrix ring \(T\) is strongly lifting.

Our first Theorem is motivated by the above-mentioned results and characterizes the semiregularity of formal triangular matrix rings relative to an ideal.

Theorem 1. Let \(T^{\prime}\) be an ideal of \(T\). Then \(T\) is \(T^{\prime}\)-semiregular if and only if \(A\) is \(I\)-semiregular, \(B\) is \(K\)-semiregular and \(N=M\).

Proof. First recall the fact that \(T\) is \(T^{\prime}\)-semiregular if and only if \(T / T^{\prime}\) is regular and \(T^{\prime}\) is strongly lifting. Now if \(T\) is \(T^{\prime}\)-semiregular, then \(T / T^{\prime}\) is regular, and so \(A / I\) and \(B / K\) regular. Further, \(\mathrm{J}\left(T / T^{\prime}\right)=0\) implies that \(M / N=0\), that is \(N=M\). Moreover, Corollary 2.8 in 1 states that strongly lifting ideals \(T^{\prime}\) of \(T\) are those ideals for which \(I\) and \(K\) are strongly lifting in \(A\) and \(B\), respectively. Combining
these two results with the above-mentioned fact, we get \(A\) is \(I\)-semiregular, \(B\) is \(K\)-semiregular, as desired.

For the converse, first note that \(T / T^{\prime}=\left(\begin{array}{cc}A / I & 0 \\ 0 & B / K\end{array}\right) \cong A / I \times B / K\). Hence, the regularity of \(A / I\) and \(B / K\) implies the regularity of \(T / T^{\prime}\). Now, the result is easily seen by again using Corollary 2.8 in 1 .

As an application, we continue with an illustrative example.
Example 1. Let \(A=\mathbb{Z}_{30}, B=\mathbb{Z}_{9}\) and \(M=\mathbb{Z}_{3} \oplus \mathbb{Z}_{3}\). Let us begin by considering the following ring:
\[
T=\left(\begin{array}{cc}
\mathbb{Z}_{30} & 0 \\
\mathbb{Z}_{3} \oplus \mathbb{Z}_{3} & \mathbb{Z}_{9}
\end{array}\right)
\]

It is our aim to determine all ideals \(T^{\prime}\) of \(T\) with the property that \(T\) is \(T^{\prime}\) semiregular by using Theorem 1 To do this, we first need to specify the strongly lifting ideals \(I(K)\) of \(A(B)\) for which \(A / I(B / K)\) is von Neumann regular, respectively. Since \(A\) and \(B\) are exchange rings, all ideals of these two rings are strongly lifting and an easy computation shows that all factors of the form \(A / I\) and \(B / K\) are von Neumann regular except for the case \(K=0\) in \(B\).

Taking into account the ideal structure of \(T\) described in Proposition 1 and letting \(N=M\), the following ideals \(T^{\prime}\) are those for which \(T\) is \(T^{\prime}\)-semiregular:
\[
\begin{aligned}
& \bullet\left(\begin{array}{cc}
0 & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
15 \mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
10 \mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
6 \mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right) ; \\
& \bullet\left(\begin{array}{cc}
5 \mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
3 \mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
2 \mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
\mathbb{Z}_{30} & 0 \\
M & \mathbb{Z}_{9}
\end{array}\right) ; \\
& \bullet\left(\begin{array}{cc}
0 & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
15 \mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
10 \mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
6 \mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right) ; \\
& \bullet\left(\begin{array}{cc}
5 \mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
3 \mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
2 \mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right),\left(\begin{array}{cc}
\mathbb{Z}_{30} & 0 \\
M & 3 \mathbb{Z}_{9}
\end{array}\right) \text {. }
\end{aligned}
\]

Remark 1. Note that the "if" part of the above theorem is not in general true if we omit the condition \(N=M\) as shown in the following example.

Let \(A=\mathbb{Z}_{4}, B=\mathbb{Z}_{2}\) and \(M=\mathbb{Z}_{2} \oplus \mathbb{Z}_{2}\). Consider the formal triangular matrix ring
\[
T=\left(\begin{array}{cc}
\mathbb{Z}_{4} & 0 \\
\mathbb{Z}_{2} \oplus \mathbb{Z}_{2} & \mathbb{Z}_{2}
\end{array}\right)
\]

We wish to find an ideal \(T^{\prime}\) of \(T\) for which \(A\) is \(I\)-semiregular, \(B\) is \(K\)-semiregular, but \(T\) is not \(T^{\prime}\)-semiregular due to the fact that \(N \neq M\). For this, we first observe that the ideals \(I=2 \mathbb{Z}_{4}\) of \(\mathbb{Z}_{4}\) and \(K=0\) of \(\mathbb{Z}_{2}\) are strongly lifting, respectively. Further, \(A / I \cong B / K \cong \mathbb{Z}_{2}\) is clearly von Neumann regular. Hence, these two together imply that \(A\) is \(I\)-semiregular and \(B\) is \(K\)-semiregular.

Taking into account the ideal structure of \(T\) described in Proposition 1 we let \(N=\mathbb{Z}_{2} \oplus 0\) a \((B, A)\)-subbimodule of \(M\) and
\[
T^{\prime}=\left(\begin{array}{cc}
2 \mathbb{Z}_{4} & 0 \\
\mathbb{Z}_{2} \oplus 0 & 0
\end{array}\right) .
\]

Then the ring \(T\) is not \(T^{\prime}\)-semiregular since the ring
\[
T / T^{\prime} \cong\left(\begin{array}{ll}
\mathbb{Z}_{2} & 0 \\
\mathbb{Z}_{2} & \mathbb{Z}_{2}
\end{array}\right)
\]
is not von Neumann regular.

It should be noted that the Jacobson radical \(\mathrm{J}(R)\) of a ring \(R\) is not necessarily idempotent lifting. However, if it is idempotent lifting, it is also strongly lifting. In fact, as is well-known, a ring \(R\) is semiregular if and only if \(R / \mathrm{J}(R)\) is regular and \(\mathrm{J}(R)\) is idempotent lifting. Hence, the \(\mathrm{J}(R)\)-semiregular rings are just the semiregular rings and we get the following immediate corollary to the above result.

Corollary 1. \(T\) is semiregular if and only if \(A\) and \(B\) are semiregular.
A right \(I\)-semiperfect ring is one in which every right ideal \(M\) of \(R\) fulfills the equivalent conditions as stated in Lemma 1. Left \(I\)-semiperfect rings can be defined in a similar vein. A short proof of the right-left symmetry of this notion appears in 9 by showing the following equivalence
\(R\) is \(I\)-semiperfect \(\Leftrightarrow R / I\) is semisimple and \(I\) is strongly lifting.
We now determine a necessary and sufficient condition for the triangular matrix ring \(T\) to be \(T^{\prime}\)-semiperfect.

Theorem 2. If \(T^{\prime}\) is an ideal of \(T\), then the following conditions are equivalent:
(i) \(T\) is \(T^{\prime}\)-semiperfect;
(ii) \(A\) is I-semiperfect, \(B\) is \(K\)-semiperfect and \(N=M\).

Proof. To begin with, let us recall that \(T\) is \(T^{\prime}\)-semiperfect if and only if \(T / T^{\prime}\) is semisimple and \(T^{\prime}\) is strongly lifting. Now if \(T\) is \(T^{\prime}\)-semiperfect, then \(T / T^{\prime}\) is semisimple, and so are \(A / I\) and \(B / K\). Further, \(J\left(T / T^{\prime}\right)=0\) implies that \(M / N=0\), that is \(N=M\). Moreover, Corollary 2.8 in 1 states that strongly lifting ideals \(T^{\prime}\) of \(T\) correspond to those ideals for which \(I\) and \(K\) are strongly lifting in \(A\) and \(B\). Combining these two results with the above-mentioned fact, we get \(A\) is \(I\)-semiperfect, \(B\) is \(K\)-semiperfect, as desired.

For the converse, consider the ring isomorphism \(T / T^{\prime}=\left(\begin{array}{cc}A / I & 0 \\ 0 & B / K\end{array}\right) \cong A / I \times\) \(B / K\). Hence, the semisimplicity of \(A / I\) and \(B / K\) implies the semisimplicity of \(T / T^{\prime}\). We get our assertion by putting these and Corollary 2.8 in 1 together.

Example 2. Let \(\mathbb{Z}_{(p)}\) be the localization of the ring of integers \(\mathbb{Z}\) at a prime ideal \(p \mathbb{Z}, \mathbb{Z}_{p \infty}\) be the Prüfer group and let \(\hat{\mathbb{Z}}_{p}\) be \(p\)-adic integers.

Consider the formal triangular matrix ring
\[
T=\left(\begin{array}{ll}
\mathbb{Z}_{(p)} & 0 \\
\mathbb{Z}_{p \infty} & \hat{\mathbb{Z}}_{p}
\end{array}\right)
\]

Our goal is to determine all ideals \(T^{\prime}\) of \(T\) with the property that \(T\) is \(T^{\prime}\) semiperfect by using Theorem 2, For this, it is enough to identify all strongly lifting ideals of \(\mathbb{Z}_{(p)}\) and \(\hat{\mathbb{Z}}_{p}\) for which every factor ring of these two rings is semisimple Artinian. Since \(\mathbb{Z}_{(p)}\) and \(\hat{\mathbb{Z}}_{p}\) are exchange rings, all ideals of these two rings are strongly lifting. It is not difficult to see that factor rings are semisimple Artinian for the ideals \(p \mathbb{Z}_{(p)}\) and \(\mathbb{Z}_{(p)}\) and the ideals \(p \hat{\mathbb{Z}}_{p}\) and \(\hat{\mathbb{Z}}_{p}\) of the uniserial rings \(\mathbb{Z}_{(p)}\) and \(\hat{\mathbb{Z}}_{p}\), respectively.

Hence, the ring \(T\) is \(T^{\prime}\)-semiperfect for the following list of ideals \(T^{\prime}\) :
\[
\left(\begin{array}{cc}
p \mathbb{Z}_{(p)} & 0 \\
\mathbb{Z}_{p^{\infty}} & p \hat{\mathbb{Z}}_{p}
\end{array}\right),\left(\begin{array}{cc}
p \mathbb{Z}_{(p)} & 0 \\
\mathbb{Z}_{p^{\infty}} & \hat{\mathbb{Z}}_{p}
\end{array}\right),\left(\begin{array}{cc}
\mathbb{Z}_{(p)} & 0 \\
\mathbb{Z}_{p^{\infty}} & p \hat{\mathbb{Z}}_{p}
\end{array}\right),\left(\begin{array}{cc}
\mathbb{Z}_{(p)} & 0 \\
\mathbb{Z}_{p^{\infty}} & \hat{\mathbb{Z}}_{p}
\end{array}\right)
\]

In the example below, it can be seen that the assumption" \(N=M\) " in the "if" part of Theorem 2 cannot simply be dropped.

Example 3. As an example of an ideal \(T^{\prime}\) of \(T\) for which \(A\) is \(I\)-semiperfect, \(B\) is \(K\)-semiperfect, but \(T\) is not \(T^{\prime}\)-semiperfect, we would like to recall an example of Berberian that was discussed in detail in 2, Example 1].

Let \(\mathbb{C}\) be the complex field and \(\mathbb{H}=\mathbb{R}+\mathbb{R} i+\mathbb{R} j+\mathbb{R} k\) be the division ring of real quaternions.

Take \(A=\mathbb{C}, B=\mathbb{H}\) and \(M=\mathbb{H}\). Let us begin by considering the following ring
\[
T=\left(\begin{array}{cc}
\mathbb{C} & 0 \\
\mathbb{H} & \mathbb{H}
\end{array}\right)
\]

The ring \(T\) is an exchange ring due to the fact that the rings \(\mathbb{C}\) and \(\mathbb{H}\) are all exchange rings 6, Proposition 2.1].

We further observe that the ideals \(I=0\) of \(\mathbb{C}\) and \(K=\mathbb{H}\) of \(\mathbb{H}\) are strongly lifting as exchange rings are precisely the rings that every one-sided ideal is strongly lifting. Furthermore, \(A / I\) and \(B / K\) are clearly semisimple. Hence, we have \(A\) is \(I\)-semiperfect and \(B\) is \(K\)-semiperfect.

On the other hand, for the ideal
\[
T^{\prime}=\left(\begin{array}{cc}
0 & 0 \\
0 & \mathbb{H}
\end{array}\right)
\]
of the ring \(T, T\) is not \(T^{\prime}\)-semiperfect since the ring
\[
T / T^{\prime} \cong\left(\begin{array}{ll}
\mathbb{C} & 0 \\
\mathbb{H} & 0
\end{array}\right)
\]
is not semisimple Artinian.

A ring \(R\) is called semipotent if each right ideal of \(R\) that is not contained in its Jacobson radical \(J(R)\) contains a nonzero idempotent. Note that this notion is left-right symmetric. A semipotent ring \(R\) is called potent if, in addition, \(\mathrm{J}(R)\) is an idempotent lifting ideal of \(R\). Examples of these rings include exchange rings (see [7, Proposition 1.9]). It is well known that a formal triangular matrix ring \(T\) is semipotent (respectively, potent) if and only if \(A\) and \(B\) are semipotent (respectively, potent) 4, Theorem 6.4].

Semipotent rings has been generalized to semipotent rings relative to an ideal based on the following lemma proposed by Nicholson and Zhou 9.

Lemma 2. [9, Lemma 19] The following are equivalent for \(I \triangleleft R\) :
(1) If \(I^{\prime} \nsubseteq I\) is a right ideal, then there exists \(e^{2}=e \in I^{\prime}-I\).
(2) If \(a \notin I\), then there exists \(e^{2}=e \in a R-I\).
(3) If \(a \notin I\) there exists \(x \in R\) such that \(x a x=x \notin I\).

Following Nicholson and Zhou, for an ideal \(I\) in a ring \(R, R\) is said to be \(I\) semipotent if the above conditions in Lemma 2 are fulfilled, and is said to be \(I\) potent if it is \(I\)-semipotent and \(I\) is strongly lifting in \(R\). In other words, the semipotent (potent) rings are simply the \(\mathrm{J}(R)\)-semipotent \((\mathrm{J}(R)\)-potent) rings. Since the property of being a semipotent or a potent ring transfers to formal triangular matrix rings by the above-mentioned result due to Haghany and Varadarajan 4. Theorem 6.4], it is natural to suspect that it may also transfer in the relative case.

We now interpret this notion in the language of formal triangular matrix rings. We mimic the proof of Haghany and Varadarajan 4, Theorem 6.4].

Theorem 3. Let \(T^{\prime}\) be an ideal of \(T\). If \(T\) is \(T^{\prime}\)-semipotent then \(A\) is I-semipotent and \(B\) is \(K\)-semipotent, respectively.

Proof. Assume that \(T\) is \(T^{\prime}\)-semipotent. We first claim that \(A\) is \(I\)-semipotent. Let \(I^{\prime} \nsubseteq I\) be a right ideal in \(A\). Then \(I^{\prime \prime}=\left(\begin{array}{cc}I^{\prime} & 0 \\ 0 & 0\end{array}\right)\) is a right ideal not contained in \(T^{\prime}\). Hence there exists \(e \in I^{\prime}\) with \(\left(\begin{array}{cc}e & 0 \\ 0 & 0\end{array}\right)\left(\begin{array}{cc}e & 0 \\ 0 & 0\end{array}\right)=\left(\begin{array}{ll}e & 0 \\ 0 & 0\end{array}\right) \in I^{\prime \prime}-T^{\prime}\). This implies that \(e^{2}=e \in I^{\prime}-I\), as desired. Secondly, we claim that \(B\) is \(K\)-semipotent. Let \(K^{\prime} \nsubseteq K\) be a right ideal in \(B\). Then \(K^{\prime \prime}=\left(\begin{array}{cc}0 & 0 \\ K^{\prime} M & K^{\prime}\end{array}\right)\) is a right ideal of \(T\) not contained in \(T^{\prime}\). Since \(T\) is \(T^{\prime}\)-semipotent, there exists a nonzero element \(\left(\begin{array}{cc}0 & 0 \\ m & f\end{array}\right) \in K^{\prime \prime}-T^{\prime}\) with \(\left(\begin{array}{cc}0 & 0 \\ m & f\end{array}\right)\left(\begin{array}{cc}0 & 0 \\ m & f\end{array}\right)=\left(\begin{array}{cc}0 & 0 \\ m & f\end{array}\right) \in K^{\prime \prime}-T^{\prime}\). This implies that \(f^{2}=f\) and \(f m=m\). Since \(\left(\begin{array}{cc}0 & 0 \\ m & f\end{array}\right)\) is nonzero, we get \(f \neq 0\) or \(m \neq 0\). By considering \(f m=m\), we get \(f \neq 0\). Thus, \(0 \neq f\) with \(f^{2}=f \in K^{\prime}-K\), as desired.

The converse of Theorem 3 does not hold in general, as can be seen in the following example.

Example 4. There exist a formal triangular ring \(T\) and an ideal \(T^{\prime}\) of \(T\) such that \(A\) is \(I\)-semipotent, \(B\) is \(K\)-semipotent, but \(T\) is not \(T^{\prime}\)-semipotent.

Let \(R=\mathbb{Z} \times \mathbb{Z}_{2} \times \mathbb{Z}_{2} \times \cdots\) be the direct product of rings. Take into consideration the following subring of \(R\) :
\[
A=\left\{\left(n, \bar{n}_{2}, \bar{n}_{3}, \cdots, \bar{n}_{k}, \bar{n}, \cdots\right) \mid n, n_{i} \in \mathbb{Z}, k \geqslant 2\right\}
\]

Putting \(I=\{(2 m, \overline{0}, \overline{0}, \cdots) \mid m \in \mathbb{Z}\}\), it follows that \(A\) is \(I\)-semipotent by 9 , Example 23].

Now, take \(B=\mathbb{Z}_{4}\) and \(M=\mathbb{Z}_{2} \oplus \mathbb{Z}_{2}\). Consider the formal triangular matrix ring
\[
T=\left(\begin{array}{cc}
A & 0 \\
M & B
\end{array}\right)
\]

We further consider the ideal \(K=2 \mathbb{Z}_{4}\) of \(B=\mathbb{Z}_{4}\). Since \(B\) is \(K\)-semiregular by Remark [1] is \(K\)-semipotent, too.

On the other hand, for the ideal
\[
T^{\prime}=\left(\begin{array}{cc}
I & 0 \\
\mathbb{Z}_{2} \oplus 0 & K
\end{array}\right)
\]
of the ring \(T, T\) is not \(T^{\prime}\)-semipotent. To show the last statement, consider the following ideal of the ring \(T\)
\[
\tilde{T}=\left(\begin{array}{cc}
I & 0 \\
0 \oplus \mathbb{Z}_{2} & K
\end{array}\right)
\]

Then \(\tilde{T}\) is clearly not contained in the ideal \(T\). On the other hand, an easy computation shows that the ideal \(\tilde{T}\) of \(T\) doesn't contain any nonzero idempotent. Thus, there do not exist any idempotent in \(\tilde{T}\) which is not in \(T^{\prime}\). By Lemma 2 \(T\) is not \(T^{\prime}\)-semipotent, as desired.

As we mentioned above, relative potent rings is a proper subclass of the class of relative semipotent rings with the additional strongly lifting condition on the relative ideal. Due to the fact that strongly lifting ideals \(T^{\prime}\) of \(T\) are those ideals for which \(I\) and \(K\) are strongly lifting in \(A\) and \(B\), respectively 1, Corollary 2.8], Theorem 3 implies the the following immediate corollary.
Corollary 2. Let \(T^{\prime}\) be an ideal of \(T\). If \(T\) is \(T^{\prime}\)-potent, then \(A\) is I-potent and \(B\) is \(K\)-potent, respectively.

The converse of Theorem 3 is not true in general, it is natural to ask the question what additional conditions are required for this to happen. We will show below that this question has an affirmative answer for ideals of the form \(\left(\begin{array}{cc}I & 0 \\ M & K\end{array}\right)\) in the formal triangular matrix ring \(T\).

Theorem 4. Let \(T^{\prime}\) be an ideal of the form \(\left(\begin{array}{cc}I & 0 \\ M & K\end{array}\right)\) in \(T\). If A is I-semipotent and \(B\) is \(K\)-semipotent, then \(T\) is \(T^{\prime}\)-semipotent.

Proof. Assume that \(A\) is \(I\)-semipotent and \(B\) is \(K\)-semipotent. We will show that \(T\) is \(T^{\prime}\)-semipotent. Let \(\tilde{T}\) be a right ideal of \(T\) with \(\tilde{T} \nsubseteq T^{\prime}\). Then there exists an element \(t=\left(\begin{array}{cc}a & 0 \\ m & b\end{array}\right) \in \tilde{T}-T^{\prime}\). This implies that either \(a \notin I\) or \(b \notin K\). First, assume that \(a \notin I\). Then there exist a non-zero idempotent \(e\) in \(a A-I\). Set \(e=a r\) for some \(r \in A\). Then are \(=e^{2}=e \neq 0\). Note that
\[
\left(\begin{array}{cc}
e & 0 \\
m r e & 0
\end{array}\right)=\left(\begin{array}{cc}
\text { are } & 0 \\
m r e & 0
\end{array}\right)=\left(\begin{array}{ll}
a & 0 \\
m & b
\end{array}\right)\left(\begin{array}{cc}
r e & 0 \\
0 & 0
\end{array}\right),
\]
and so \(\left(\begin{array}{cc}e & 0 \\ m r e & 0\end{array}\right) \in t T \subseteq \tilde{T}\). Also
\[
\left(\begin{array}{cc}
e & 0 \\
m r e & 0
\end{array}\right)\left(\begin{array}{cc}
e & 0 \\
m r e & 0
\end{array}\right)=\left(\begin{array}{cc}
e & 0 \\
m r e & 0
\end{array}\right) .
\]

Thus \(\left(\begin{array}{cc}e & 0 \\ m r e & 0\end{array}\right)\) is a non-zero idempotent in \(t T-T^{\prime}\). If \(b \notin K\), there exists a non-zero idempotent \(f \in b B\) and \(\left(\begin{array}{ll}0 & 0 \\ 0 & f\end{array}\right)\) is a non-zero idempotent in \(t T-T^{\prime}\). This proves that \(T\) is \(T^{\prime}\)-semipotent.

Considering the definition of a relative potent (resp. semipotent and potent) ring we end the paper with the following corollaries of Theorem 3 and Theorem 4

Corollary 3. Let \(T^{\prime}\) be an ideal of of the form \(\left(\begin{array}{cc}I & 0 \\ M\end{array}\right)\) in \(T\). If \(A\) is I-potent and \(B\) is \(K\)-potent, then \(T\) is \(T^{\prime}\)-potent.

Corollary 4. 4. Theorem 6.4] Let \(T\) be the formal triangular matrix ring of the form ( \(\begin{gathered}A \\ M\end{gathered} 0_{B}\) ). Then \(T\) is semipotent (resp. potent) if and only if \(A\) and \(B\) are semipotent (resp. potent).

Declaration of Competing Interests The author declares that there is no competing interest regarding the publication of this paper.
Acknowledgements The author would like to thank the referees for their careful considerations.

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http://communications.science.ankara.edu.tr

\title{
MULTIGRID METHODS FOR NON COERCIVE VARIATIONAL INEQUALITIES
}

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\begin{abstract}
In this study, our examination centers around the numerical resolution of non-coercive issues using a multi-grid approach. Our particular emphasis is directed towards employing multi-grid methodologies to tackle non-linear variational inequalities. Our primary goal involves confirming the consistent convergence of the multi-grid algorithm. To attain this objective, we make use of fundamental sub-differential calculus and glean insights from the convergence principles of non-linear multi-grid techniques.
\end{abstract}

\section*{1. Introduction}

Contemporary literature showcases a diverse array of computational technique that are harnessed to address intricate real-world challenges spanning various scientific and engineering domains. These methodologies have been crafted and utilized to confront demanding problems, yielding efficient resolutions within their respective fields. Many researchers have explored these computational strategies to tackle a number of applied problems, propelling comprehension and advance understanding and progress in many scientific fields.
Commonly used numerical methods for solving boundary problems generally lead, after discretisation, to the solution of systems of algebraic equations. These numerical techniques, encompassing iterative methods like Jacobi, Gauss-Seidel iteration, and relaxation methods, are frequently chosen due to their conventional nature.However, they may show a slow convergence of fine mesh sizes and complexity when applied to general ellipticity problems. In contrast, multi-grid methods offer

\footnotetext{
2020 Mathematics Subject Classification. 65M55, 65N30, 49N05, 65K15, 35J86.
Keywords. Multigrid method, variational inequality, finite element, iterative method, HJB equation, non coercive operator.

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a clear advantage. These algorithms exhibit linear expenses based on the number of discretization points. These algorithms exhibit linear expenses based on the number of discretization points, regardless of the problem's dimensions. Particularly, these methods are adept at resolving linear and non-linear partial differential equations (PDEs) as well as linear V.Is (Variational inequalities) [12, 10, 7. Their linear complexity makes them powerful tools for large problems, greatly reducing computational requirements while ensuring accurate solutions. Multi-grid techniques are widely praised as a fast approach to tackling various forms of variational equations and inequalities [11], particularly in the area the discretized elliptic problems that leads to an \(M\)-matrix [6].
Through a conforming finite element method \(P_{1}\) [4], we will be providing an overview of non-linear variational inequalities (N.V.I) problems and their discretization in the following section. Additionally, The Hoppe multi-grid method [14, 9] served as an inspiration for our algorithm, which views the V.I as stationary Hamilton-JacobiBellman(H.J.B) equations. The iteration matrices are provided for an algorithm known as the M.G.H.J.B, or multi-grid Hierarchy Jacobi.
First, we present original results on the approximation and smoothness properties within the \(L^{\infty}\) norm. We then demonstrate the consistent convergence of the M.G.H.J.B algorithm. Finally, we apply the numerical method to a specific scenario where the operator is linear and unconstrained, and the second element is independent of the solution. In this context, we implemented the Gauss-Seidel method and the multigrid method V and W cycles. Numerical experiments are performed to evaluate the efficiency and performance of these methods in solving the proposed problem.

\section*{2. Multigrid Method}
2.1. Assumptions and Notations. Suppose that \(\Omega\) is an open in \(\mathbb{R}^{N}\) with a sufficiently regular border \(\partial \Omega\).
We define second order operators with \(u, v \in H^{1}(\Omega)\),
\[
\mathfrak{A}=\sum_{1 \leq j, k \leq N} \partial_{j k}(x) \frac{\partial^{2}}{\partial x_{j} \partial x_{k}}+\sum_{k=1}^{N} \mathfrak{b}_{k}(x) \frac{\partial}{\partial x_{k}}+\mathfrak{b}_{0}(x),
\]
where \(\partial_{j k}(x), \mathfrak{b}_{k}(x), \mathfrak{b}_{0}(x)\) are sufficiently regular coefficients such that:
\[
\partial_{k j}(x)=\partial_{j k}(x), \quad \mathfrak{b}_{0}(x) \geq \beta>0 ; \quad(x \in \Omega)
\]

Also, we define the associated bilinear non-coercive forms
\[
\mathfrak{a}(u, v)=\int_{\Omega}\left(\sum_{1 \leq j, k \leq N} \partial_{j k} \frac{\partial u}{\partial x_{j}} \frac{\partial v}{\partial x_{k}}+\sum_{k=1}^{N} \mathfrak{b}_{k} \frac{\partial u}{\partial x_{k}} v+\mathfrak{b}_{0}(x) u v\right) d x
\]
and the operators
\[
\begin{equation*}
\mathcal{B}=\sum_{1 \leq j, k \leq N} \partial_{j k}(x) \frac{\partial^{2}}{\partial x_{j} \partial x_{k}}+\sum_{k=1}^{N} \mathfrak{b}_{k}(x) \frac{\partial}{\partial x_{k}}+\left(\mathfrak{b}_{0}(x)+\lambda\right) \tag{1}
\end{equation*}
\]
we choose \(\lambda>0\) is sufficiently large so that \(\mathcal{B}=\mathfrak{A}+\lambda I\) are strongly elliptic on \(H^{1}(\Omega)\) and
\[
\begin{equation*}
\mathfrak{b}(u, v)=\mathfrak{a}(u, v)+\lambda(u, v) \tag{2}
\end{equation*}
\]

Additionally, we consider \(f\) a second member as following:
\[
f \in L^{\infty}(\Omega) ; \quad f \geq 0
\]
and obstacle \(\psi \in W^{2, \infty}\), where \(\psi>0\).
2.2. Problem Continuous. The aim is to find \(u\) the solution of the problem presented by the following V.Is:
Find \(u\) solution of:
\[
\left\{\begin{array}{l}
\mathfrak{b}(u, v-u) \geq(f+\lambda u, v-u), \quad \forall v \in H^{1}(\Omega),  \tag{3}\\
u \leq \psi ; \quad v \leq \psi
\end{array}\right.
\]

It has been confirmed that this issue has a singular solution, as demonstrated by the theorem of fixed point and from the aforementioned assumptions (see [1]).
2.3. Discretization. In order to build a multi-grid loop, we create a sequence of discretization steps referred to as \(0<\mathfrak{h}_{k+1}<\mathfrak{h}_{k}<1\) such that the grids are nested \(\mathfrak{h}_{k+1}=\frac{\mathfrak{h}_{k}}{2}\).
Subsequently, we delineate \(\Omega_{k}=\Omega_{\mathfrak{h}_{k}}, V_{k}=V_{\mathfrak{h}_{k}}, \mathfrak{A}_{k}=\mathfrak{A}_{\mathfrak{h}_{k}}\) and we establish a series of uniform regular triangulations referred to as \(\left\{T_{k}, k \in \mathbb{N}_{0}\right\}\). For all \(T_{k}\), we have
\[
\begin{aligned}
& \Omega_{k} \subset \Omega_{k+1} \subset \Omega \\
& \operatorname{dist}\left(\partial \Omega_{k}, \partial \Omega\right) \leq c_{0} \mathfrak{h}_{k}^{2} \\
& \mathfrak{h}_{k} \mathfrak{h}_{k+1} \leq c_{1}
\end{aligned}
\]

We introduce \(V_{\mathfrak{h}_{k}}=\left\{v_{\mathfrak{h}_{k}} \in C(\Omega) \cap H^{1} ; v_{\mathfrak{h}_{k}} / T \in P_{1}\right\}\), for simplicity we write:
\[
V_{k}=\left\{v_{k} \in C(\Omega) \cap H^{1} ; \quad v_{k} / r \in P_{1}\right\}
\]

The shape function \(\varphi_{k}^{i}, i \in\left(1, \ldots, m\left(\mathfrak{h}_{k}\right)\right)\) of the usual basis is defined as: \(\varphi_{k}^{i}\left(x_{k}^{j}\right)=\) \(\delta_{i j}\), where \(x_{k}^{j}\) be a node of the \(T_{k}\) triangulation .
So, the ordinary restriction operator \(r_{k}\) is defined like:
\[
\begin{equation*}
r_{k} v(x)=\sum_{i=1}^{m\left(\mathfrak{h}_{k}\right)} v\left(M_{k}^{i}\right) \varphi_{k}^{i}(x) . \tag{4}
\end{equation*}
\]

If we suppose \(U_{k}=\mathbb{R}^{m_{k}}\). Then, \(r_{k}: U_{k} \rightarrow V_{k}\) is a bijection.
\(U_{k}\) is equipped with the scalar product
\[
<u, v>=\mathfrak{h}_{k}^{2} \sum_{i=1}^{m\left(\mathfrak{h}_{k}\right)} u_{i} v_{i}, \quad\|u\|_{k}=<u, v>_{k}^{1 / 2}
\]

The maximum norms in \(U_{k}\) and \(V_{k}\) are equivalent, we denote them \(\|\cdot\|_{\infty}\). We have the following lemma (see [2]).

Lemma 1. There exists \(C_{1}, C_{2}\) independent of \(k\) such that
\[
\begin{array}{ll}
\left\|r_{k}(u)\right\|_{\infty} & =\|u\|_{\infty}, \quad \forall u \in U_{k} \\
C_{1}\|v\|_{\infty} & \leq\left\|r_{k}^{*}(v)\right\|_{\infty} \leq C_{2}\|v\|_{\infty}, \quad \forall v \in V_{k} \tag{5}
\end{array}
\]
2.4. Problem Discrete. Continuing in a logical sequence, we present the discretization matrices \(\mathcal{B}_{k}\) and the bilinear form \(b\left(\varphi_{k}^{1}, \varphi_{k}^{s}\right)\), where \(\varphi_{s}\) the shape functions. With these descriptions established. Now, we are positioned to formulate the discrete problem in the subsequent manner:
Find \(u_{k} \in V_{k}\) solution of:
\[
\left\{\begin{array}{l}
<\mathcal{B}_{k} u_{k}, v_{k}-u_{k}>\geq<f_{k}+\lambda u_{k}, v_{k}-u_{k}>, \quad \forall v_{k} \in V_{k}  \tag{6}\\
u_{k} \leq r_{k} \psi, \quad v_{k} \leq r_{k} \psi
\end{array}\right.
\]

We make the assumption that the matrices \(\mathcal{B}_{k}\) are \(M\)-matrices.(see [3]).
2.5. H.J.B form. The correspondence between the finite-dimensional V.I (3) and a representation in Hamilton-Jacobi-Bellman (H.J.B) form is easily discernible (see [10]). We detail the selected numerical technique for resolving the stationary H.J.B equations.
In the traditional framework, we recollect certain convergence outcomes that will play a crucial role in affirming the M.G.H.J.B algorithm's convergence expounded in the following:
Iterative diagram:
Step 1: Choose \(u_{k}^{0} \in \mathbb{R}^{n_{k}}\) as initial vector.
Step 2 : Calculate the solution \(u_{k}^{\nu+1} \in \mathbb{R}^{n_{k}}\) of the following recurrence equation
\[
\begin{equation*}
\mathcal{B}_{k}^{\nu} u_{k}^{\nu+1}-Z_{k}^{\nu}=0 \tag{7}
\end{equation*}
\]
such that
\[
Z_{k}^{\nu}=F_{k}^{\nu}+\lambda u_{k}^{\nu}
\]
where
\[
\begin{gather*}
\mathcal{B}_{k, i}^{\nu}=\left\{\begin{array}{ccc}
\mathcal{B}_{k, i}\left(u_{k}\right) & \text { if } & \mathcal{B}_{k, i} u_{k, i}^{\nu}-Z_{k, i}>u_{k, i}^{\nu}-\psi_{k, i}, \\
u_{k, i} & \text { if } & 1 \leq i \leq N
\end{array}\right.  \tag{8}\\
Z_{k, i}^{\nu}=\left\{\begin{array}{ccc}
Z_{k, i} & \text { if } & \mathcal{B}_{k, i} u_{k, i}^{\nu}-Z_{k, i}>u_{k, i}^{\nu}-\psi_{k, i} \\
u_{k, i} & \text { if } & 1 \leq i \leq N
\end{array}\right. \tag{9}
\end{gather*}
\]

Let the discrete H.J.B equation where \(u_{k}^{*}\) be the unique solution
\[
\begin{equation*}
\max _{1 \leq i \leq N}\left(\mathcal{B}_{k, i} u_{k}^{*}-Z_{k, i}, u_{k, i}^{*}-\psi_{k, i}\right)=0 \tag{10}
\end{equation*}
\]

We will formulate the subsequent theorem and introduce our problem derived from the (H.J.B) equation, drawing inspiration from Hoppe's [10].
Theorem 1. Let \(u_{k}^{\nu}\) be the solution in the iteration defined and it satisfies the H.J.B equation. Furthermore, We make that \(\mathcal{B}_{k}\) is continuously differentiable then the sequence \(\left(u_{k}^{\nu}\right)_{\nu \geq 0}\) converges and approaches \(u_{k}^{*}\).

Previously moving forward with presenting the findings, it is relevant to revisit the subsequent theorem:
Theorem 2. (see [1], [5]) If the previous notations and assumptions are satisfies. So , we have:
\[
\begin{equation*}
\left\|u-u_{k}^{*}\right\|_{\infty} \leq C \mathfrak{h}_{k}^{2}\left|\log \mathfrak{h}_{k}\right|^{2}\|g(u)\|_{\infty} \tag{11}
\end{equation*}
\]
2.6. Multi-grid ( M.G.H.J.B) algorithm for V.Is. For the multi-grid method we choose an iteration \(u_{k}^{\nu}, \nu>0\). So, we obtain \(\bar{u}_{k}^{\nu}\), by using an iterative method to solve the system (7) by \(\alpha\)
\[
\begin{equation*}
\bar{u}_{k}^{\nu}=S_{k}^{\alpha}\left(u_{k}^{\nu}\right) \tag{12}
\end{equation*}
\]
where \(S_{k}\) is the smoothing operator and \(\alpha\) is the number performed of iterations. The solution of (7) is denoted by \(u_{k}^{*}\). The error setting \(e_{k}^{\nu}=\bar{u}_{k}^{\nu}-u_{k}^{*}\), and the residual \(d_{k}^{(\nu)}=Z_{k}^{\nu}-\mathcal{B}_{k}^{\nu} \bar{u}_{k}^{\nu}\), the equation (7) can be write as
\[
\mathcal{B}_{k}^{\nu}\left(\bar{u}_{k}^{\nu}+e_{k}^{\nu}\right)=Z_{k}^{\nu}
\]

This leads to the residual equation
\[
\mathcal{B}_{k}^{\nu} e_{k}^{\nu}=Z_{k}^{\nu}-\mathcal{B}_{k}^{\nu} \bar{u}_{k}^{\nu}=d_{k}^{\nu}
\]

After the relaxation on \(\mathcal{B}_{k}^{\nu} \bar{u}_{k}^{\nu}=Z_{k}^{\nu}\) on the fine grid, the error will display a continuous nature. However, the error on the coarse grid appears to be more oscillatory, leading to the relaxation. At the \((k-1)\) level, we need to compute \(e_{k-1}^{\nu}\) for determine \(e_{k}^{\nu}\), where \(e_{k-1}^{\nu}\) is the solution of the coarse grid system
\[
\begin{equation*}
\mathcal{B}_{k-1}^{\nu} e_{k-1}^{\nu}=d_{k-1}^{\nu} \tag{13}
\end{equation*}
\]

We can interpret \(e_{k-1}^{\nu}\left(\operatorname{resp} \mathcal{B}_{k-1}^{\nu}, d_{k-1}^{\nu}\right)\) and \(e_{k}^{\nu}\left(\operatorname{resp} \mathcal{B}_{k}^{\nu}, d_{k}^{\nu}\right)\) as approximation operator at level \((k-1)\) and \((k)\) respectively. Additionally, we have \(\mathcal{R}_{k}\) the restriction operator and \(\mathcal{P}_{k}\) its reverse .
consecquently, at the ( \(k\) ) level we identify an improved iteration
\[
\begin{equation*}
u_{k}^{\nu+1}=\bar{u}_{k}^{\nu}+\mathcal{P}_{k}\left(e_{k-1}^{\nu}\right) \tag{14}
\end{equation*}
\]

Because of the nested structure, we employ the well-defined identity operator
\[
\pi: V_{k-1} \longrightarrow V_{k} ; \quad \pi v=v
\]
the operators of extension and restriction define like
\[
\begin{equation*}
\mathcal{P}_{k}=r_{k}^{-1} r_{k-1}, \quad \mathcal{R}_{k}=\mathcal{P}_{k}^{t} \tag{15}
\end{equation*}
\]
2.7. Matrix of the M.G.H.J.B Algorithm. For each iteration, The matrix of the two-grid method with \(\alpha_{1}\) pre-smoothing and \(\alpha_{2}\) post-smoothing iterations at the ( \(k\) ) level is given by
\[
\begin{equation*}
T G_{k}\left(\alpha_{1}, \alpha_{2}\right)=S_{k}^{\alpha_{2}}\left(\left(\mathcal{B}_{k}^{\nu}\right)^{-1}-\mathcal{P}_{k}\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right)\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha_{1}} \tag{16}
\end{equation*}
\]

Theorem 3. (see 13 ) The multi-grid technique embodies a linear iterative approach, with the iteration matrix referred to as \(M G_{k}\)
\[
\begin{align*}
M G_{0} & =0 \\
M G_{k} & =S_{k}^{\alpha_{2}}\left(I_{k}-\mathcal{P}_{k}\left(I_{k}-M G_{k-1}\right)\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right)\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha_{1}}  \tag{17}\\
& =T G_{k}+S_{k}^{\alpha_{2}} \mathcal{P}_{k} M G_{k-1}\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha_{1}}, \quad k=1,2, . .
\end{align*}
\]
3. Convergence of the Multi-grid algorithm in \(L^{\infty}\)-norm

This section is devoted to presenting a unified convergence analysis of multi-grid algorithm. To prove the convergence, we need the following proprieties

\subsection*{3.1. Approximation property.}

Theorem 4. (see [8] ) The matrix \(\Upsilon_{k}=\left[\left(\mathcal{B}_{k}^{\nu}\right)^{-1}-\mathcal{P}_{k}\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right]\) has the approximation property
\[
\begin{equation*}
\left\|\Upsilon_{k}\right\|_{\infty} \leq C h_{k}^{2}\left|\ln h_{k}\right|^{2} \tag{18}
\end{equation*}
\]

Proof. The proof was proposed by Arnold in [14] on Theorem 1 .
3.2. Property of Smoothing. To prove the smoothness property, we consider the decomposition \(\mathcal{B}_{k}^{\nu}=E_{k}-N_{k}\) and using the following assumptions: for all \(k\)
\[
\begin{gather*}
E_{k} \text { is regular and }\left\|E_{k}^{-1} N_{k}\right\|_{\infty} \leq 1,  \tag{19}\\
\left\|E_{k}\right\|_{\infty} \leq C h_{k}^{-2}, \text { with } \mathrm{C} \text { independent of } \mathrm{k} . \tag{20}
\end{gather*}
\]

In the process of smoothing, we utilize a relaxation method with an iterative matrix
\[
S_{k}=I_{k}-\omega E_{k}^{-1} N_{k}, \quad \omega \in(0,1)
\]

For the following theorem, the concept of Arnold Reusken [14] is relevant to our work.

Theorem 5. Under the previous assumptions, there exists a constant \(C\), which is independent of both \(k\) and \(\alpha\). Such that:
\[
\begin{equation*}
\left\|\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha}\right\|_{\infty} \leq C \frac{1}{\sqrt{\alpha}} h_{k}^{-2} \tag{21}
\end{equation*}
\]
(smoothness properties)

By switching to the norm in (14), from (18) and (21) we can proving the following estimation:
\[
\begin{equation*}
\exists C_{s}:\left\|S_{k}^{\alpha}\right\|_{\infty} \leq C_{s}, \text { for all } k \text { and } \alpha \tag{22}
\end{equation*}
\]

From the equation (16) with two lattices iterate (two-grid) and \(\alpha_{2}=0\), we have the following estimate:
\[
\begin{aligned}
\left\|T G_{k}\left(\alpha_{1}, 0\right)\right\|_{\infty} & =\left\|\left(\left(\mathcal{B}_{k}^{\nu}\right)^{-1}-\mathcal{P}_{k}\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right)\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha_{1}}\right\|_{\infty} \\
& \leq\left\|\left(\left(\mathcal{B}_{k}^{\nu}\right)^{-1}-\mathcal{P}_{k}\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right)\right\|_{\infty}\left\|\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha_{1}}\right\|_{\infty}
\end{aligned}
\]

Typically, we choose a hierarchy of more than two-grids. in this case, we can define the iterative matrices (17) by the recurrence of (16) for all \((k)\) levels.

Theorem 6. (13) Consider a multi-grid method for a given iterative matrix (17). Then under the previous assumption, for the parameter value \(\alpha_{2}=0, \alpha_{1}=\) \(\alpha>0, \tau \geq 2\). For each \(\zeta \in(0,1)\) there is a \(\alpha^{*}\) such that for all \(\alpha \geq \alpha^{*}\)
\[
\begin{equation*}
\left\|M G_{k}\right\|_{\infty} \leq \zeta, k=0,1, \ldots \tag{23}
\end{equation*}
\]
hold.
Proof. If the previous properties are related with (22), then we can stratify the same steps as in [ [13], Theorem 7.20].

The main result of our study was in the following theorem.
Theorem 7. For two meshes \((k)\) and \((k-1)\) and the previous given the iterated \(u_{k}^{\nu}, \nu \geq 0\) satisfy:
\[
\begin{equation*}
\left\|u_{k}^{\nu+1}-u_{k}^{*}\right\|_{\infty} \leq\left(\frac{C}{\sqrt{\alpha}}\left|\operatorname{Logh}_{k}\right|^{2}\right)\left\|u_{k}^{\nu}-u_{k}^{*}\right\|_{\infty} \tag{24}
\end{equation*}
\]

Proof. We have
\[
\begin{aligned}
\left\|u_{k}^{\nu+1}-u_{k}^{*}\right\|_{\infty} & =\left\|\left(\left(I_{k}-\mathcal{P}_{k}\left(I_{k}-M G_{k-1}\right)\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right)\left(\mathcal{B}_{k}^{\nu}\right) S_{k}^{\alpha_{1}}\right)\left(u_{k}^{\nu}-u_{k}^{*}\right)\right\|_{\infty} \\
& \leq\left\|\left(I_{k}-\mathcal{P}_{k}\left(I_{k}-M G_{k-1}\right)\left(\mathcal{B}_{k-1}^{\nu}\right)^{-1} \mathcal{R}_{k}\right)\right\|_{\infty}\left\|\mathcal{B}_{k}^{\nu} S_{k}^{\alpha_{1}}\right\|_{\infty}\left\|u_{k}^{\nu}-u_{k}^{*}\right\|_{\infty} \\
& \leq\left(\frac{C_{2}}{\sqrt{\alpha}} h_{k}^{-2}\right)\left(C_{1} h_{k}^{2}\left|\log h_{k}\right|^{2}\right)\left\|\left(u_{k}^{\nu}-u_{k}^{*}\right)\right\|_{\infty} \\
& \leq\left(\frac{C_{1} C_{2}}{\sqrt{\alpha}}\right)\left|\log h_{k}\right|^{2}\left\|u_{k}^{\nu}-u_{k}^{*}\right\|_{\infty}
\end{aligned}
\]

\section*{4. Numerical Simulation}

In this part, we applied this method to the numerical example of a non-linear variational inequality.
We suppose that the problem to be sufficiently smooth data and we apply the dynamic programming principle of Bellman, then we solve (3) as we discussed before, using the following datas:

\section*{- Mixed operator}
\[
\left\{\begin{array}{l}
\mathcal{B} u \geq f, \quad \text { in } \Omega=[0,1]^{2}  \tag{25}\\
\langle\mathcal{B} u-f, u-\psi\rangle=0, \\
u \leq \psi, \\
u=0, \quad \text { in } \partial \Omega
\end{array}\right.
\]

Where
\[
\begin{aligned}
\mathcal{B} u & =-\Delta u-0.02 \frac{\partial^{2} u}{\partial x \partial y}+0.15 \frac{\partial u}{\partial x}+0.1 \frac{\partial u}{\partial y}+(1+\lambda) u \\
f & =\sin (\pi x) \sin (2 \pi y) \sin (\pi(x+y))+\lambda u \\
\lambda & =2 \\
\psi & =0
\end{aligned}
\]

We are constrain ourselves to the discretization of finite element method with a uniform triangulation and \(P_{1}\) shape functions. For the domain, we have decretized by Matlab PDE toolbox (Matlab R2017b) for mesh generation. We solve the equation (25) by the M.G with 64 triangle and 41 nodes in the domain. This numerical illustration is performed to showcase the high efficiency of the M.G method. For the pre/post-smoothing of the M.G, we choose the Gauss-Seidel (G.S) method. The degrees of freedom chooses lower than 5 ( recursion number of M.G method). Figure 1 illustrates the convergence behaviour of the M.G solver (green and red curves of M.G (V and W cycle)) with respect to the number of iterations performed. For comparison, the convergence behavior of Gauss-Seidel (blue curves) are included.

Norm of residual obtained after 100 iterations :
\begin{tabular}{c|c|c} 
by Gauss Seidel method & by multi-grid V-cycle & by multi-grid W-cycle \\
\(4.058087199609872 \mathrm{e}^{-12}\) & \(4.440892098500626 \mathrm{e}^{-16}\) & \(4.440892098500626 \mathrm{e}^{-16}\)
\end{tabular}
We have applied the Matlab-backslash-operator(M.B.O), G.S and the M.G (V and W-cycle) are carried out on the finest grid (41 grids) and on the coarsest one ( 4 nodes) then we get the solutions in figures 2 .

Norm of residual obtained after 20 iterations :
\begin{tabular}{c|c|c} 
by Gauss Seidel method & by multi-grid V-cycle & by multi-grid W-cycle \\
0.001165086612534 & \(4.440892098500626 \mathrm{e}^{-16}\) & \(4.440892098500626 \mathrm{e}^{-16}\)
\end{tabular}


Figure 1. Comparison between the convergence of maximum residual norm by M.G and G.S.

\section*{- Simple operator}
\[
\left\{\begin{array}{l}
\mathcal{B} u \geq f, \quad \text { in } \Omega=[0,1]^{2}  \tag{26}\\
\langle\mathcal{B} u-f, u-\psi\rangle=0, \\
u \leq \psi, \\
u=0, \quad \text { in } \partial \Omega
\end{array}\right.
\]

Where
\[
\begin{aligned}
\mathcal{B} u & =-\Delta u+0.5 x \frac{\partial u}{\partial x}+0.5 y \frac{\partial u}{\partial y}+(0.045+\lambda) u \\
f & =\sin (2 \pi x) \sin (2 \pi y)+\lambda u \\
\lambda & =1 \\
\psi & =0
\end{aligned}
\]

With the same steps, we have:


Figure 2. Solution of (25) on fine grid using M.B.O, G.S, M.G V-cycle and W-cycle after 100 iterations.


Figure 3. Comparison between the convergence of maximum residual norm by M.G and G.S.

Norm of residual obtained after 100 iterations :
\begin{tabular}{c|c|c} 
by Gauss Seidel method & by multi-grid V-cycle & by multi-grid W-cycle \\
\(1.076361222374089 \mathrm{e}^{-11}\) & \(2.220446049250313 \mathrm{e}^{-16}\) & \(2.220446049250313 \mathrm{e}^{-16}\)
\end{tabular}


Figure 4. Solution of (26) on fine grid using M.B.O, G.S, M.G V-cycle and W-cycle after 100 iterations.

Norm of residual obtained after 10 iterations :
\begin{tabular}{c|c|c} 
by Gauss Seidel method & by multi-grid V-cycle & by multi-grid W-cycle \\
0.020709274936256 & \(4.884981308350689 \mathrm{e}^{-15}\) & \(2.220446049250313 \mathrm{e}^{-16}\)
\end{tabular}

Remark 1. Should we conduct more than 10 iterations, the M.G approach emerges as the optimal method.
4.1. Conclusion. Discretizing elliptic V.I. via efficient iterative solutions is the main focus of our study, employing algebraic M.G. The goal is to tackle loop domains' discretization using adaptive finite element approximation. Once discretization is complete, we successfully apply M.G to address the discrete problems at hand. Our main objective is to establish uniform convergence through our approach, and our research demonstrates the M.G's significant reduction in iteration count compared to the maximum norm method.

By means of numerical experimentation, we have constructed an example of a variational inequality. Our results indicate that the G.S. method, despite a substantial number of iterations, is unsuccessful in producing satisfactory outcomes. On the other hand, through the use of an error-damping mechanism that reduces highfrequency errors and transfers low-frequency errors to a coarser grid for alleviation, M.G. significantly enhances convergence and achieves it within a limited number of iterations. Our team recognizes the exceptional potential for further development using these methodologies.
Our numerical solution could be even more efficient and scalable if we explore the prospect of applying a parallel full M.G to surmount unconstrained elliptical inequalities. This avenue presents an interesting opportunity to cater to a broader range of problem domains.

Author Contribution Statements All authors contributed equally in the writing of this paper. All authors read and approved the final manuscript.

Declaration of Competing Interests The authors declare that they have no known competing financial interests or personal relationships that could have appeared to influence the work reported in this paper.

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\title{
A WEIGHTED GOMPERTZ-G FAMILY OF DISTRIBUTIONS FOR RELIABILITY AND LIFETIME DATA ANALYSIS
}

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\begin{abstract}
This article is set to push new boundaries with leading-edge innovations in statistical distribution for generating up-to-the-minute contemporary distributions by a mixture of the second record value of the Gompertz distribution and the classical Gompertz model (weighted Gompertz model) using T-X characterization, especially used for two-sided schemes that provide an accurate model. The quantile, ordinary, and complete moments, order statistics, probability, and moments generating functions, entropies, probability weighted moments, Lin's condition random variable, reliability in multicomponent stress strength system, reversed, and moments of residuals life and other reliability characteristics in engineering, actuarial, economics, and environmental technology were derived in their closed form. To investigate and test the flexibility, viability, tractability, and performance of the proposed Weighted Gompertz-G (WGG) generated model, the shapes of some sub-models of the WGG model were examined. The shapes of the sub-models indicated J-shapes, increasing, decreasing, and bathtub hazard rate functions. The maximum likelihood estimation of the WGG-generated model parameters was examined. An illustration with simulation and real-life data analysis indicated that the WGG-generated model provides consistently better goodness-of-fit statistics than some competitive models in the literature.
\end{abstract}

\section*{1. Introduction}

Modeling real-life data set requires a distribution that has a true reflection of the character of that data. However, to unravel the interest of some important Poisson scenarios, a parsimonious statistical distribution is required. Hence, new statistical

\footnotetext{
2020 Mathematics Subject Classification. Primary 62E10; Secondary 62E15, 62E20.
Keywords. Gompertz distribution, Poisson processes, record values, weighted Gompertz, weighted model.

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models are often introduced to harness salient factors responsive for good decision making.

Oftentimes, change-point models are characterized by abrupt behavioural structures that may be very complicated to handle by the usual classical statistical distributions. These events are but not limited to macroeconomic events characterized by abrupt increases interest rates and inflation. The abrupt behavioural shift might also be the case in extreme events like the storm and rainfall events that have ravaged some countries in recent time. The advent of the novel epidemic COVID 19 is also not exempted. Another example is the lifetime scenario that are subjected to unexpected and rapid shocks. Hence, this study is proposed to deal with such change-point by constructing an appropriate weighted generated distribution called Weighted Gompertz-G (WGG) distribution that can address the differentials. Though the method of generating new distribution is not new, using the weighted generator concept to generate new models is a new approach targeted at change-point problems. Thus, this article will use the weighted Gompertz generator approach to generate new continuous distributions that are more flexible, and viable in their goodness-of-fit test statistics.

The Gompertz model has played a vital role in modeling scenarios that deal with survival times, reliability, human mortality, and actuarial data with exponential increase outcomes. Thus, it has received considerable attention from demographers, economics, and actuaries. This includes 13 , and 12 who proposed the shifted Gompertz-G and alpha power Teissier distributions. A flexible alpha power Gompertz distribution was proposed in 14.27 emphasised on some applications of the Gompertz distribution in Poisson process. A negative rate of aging parameter with Gompertz distribution was proposed in 22.8 proposed the Teissier distribution. 16 proposed the Marshall-Olkin Teissier distribution. The gamma-Gompertz distribution was proposed in 29 . 15 developed the alpha power Marshall-Olkin-G model. 23 developed the Topp-Leone Gompertz distribution with application to glass data. The reliability properties and applications of the alpha power Topp-Leone-G distribution was considered in 17. However, some researches have been contributed to generating newer classical statistical distributions include 5 and 2 who proposed exponentiated T-X and T-X family of distributions. The type I half-logistic family of distributions proposed by 10 . The beta and generalized gamma-generated distributions by 30 . A tetration distribution developed by 11 . Odd Truncated Inverse Exponential Weibull Exponential by 11. 24 proposed a New Member from the T-X Family of distribution. A New Odd Log-Logistic Lindley Distribution was proposed in 3. The Bivariate Lack-of-Memory Distributions was developed in 21 . 20 proposed the U family of distributions. A new estended Weibull distribution was developed in 26. 18 proposed the alpha power TeissierG Distribution and its Applications in reliability analysis. Exponentiated Gumbel Weibull Logistic model was developed in 25. Weighted Weibull-G was introduced by 19 .

Let \(T\) be a nonnegative random variable with a probability density function (pdf) \(f(t)\) such that for a suppose \(t>0\), weight function \(w(t)=\beta+\exp (\beta t)-1\), and expectation \(E[w(t)]=\frac{\beta \lambda+1}{\lambda}\). Then, 7 defined the pdf and cumulative distribution function (cdf) \(F(t)\) of the weighted Gompertz distribution as
\[
\begin{equation*}
f(t)=\frac{\beta \lambda^{2}}{(1+\beta \lambda)}\left(\beta-1+e^{\beta t}\right) e^{\left(\beta t-\lambda\left(e^{\beta t}-1\right)\right.}, t>0 \beta, \lambda>0 \tag{1}
\end{equation*}
\]
and
\[
\begin{equation*}
F(t)=1-\left[1+\frac{\lambda\left(e^{\beta t}-1\right)}{(1+\beta \lambda)}\right] e^{-\lambda\left(e^{\beta t}-1\right)}, t>0 \beta, \lambda>0 \tag{2}
\end{equation*}
\]
with \(\lambda\) and \(\beta\) as the shape and scale parameters.
Modeling abrupt behavioural structure and scenarios has become more complicated as a result of their change-point. Though the method of generating new distribution is not new, using the weighted generator concept to generate new models is a new approach. Hence, this study is motivated to propose a model with a true reflection of the character of the data obtained. Thus, the WGG generated model tends to improve the goodness-of-fit, and the test statistics of the existing distributional models using weighted distribution characterization.

The study aim at introducing a class of generator with the aid of the weighted Gompertz model called the weighted Gompertz generator. This generated model will improve the performance, flexibility and the viability of the goodness-of-fit of the abrupt behaviourial change-point scenarios in lifetime modeling.

\section*{2. The Weighted Gompertz-G Distribution}

Suppose a nonnegative random variable \(T\) is defined on the interval \(T \in[m, n]\) for \(-\infty<m<n<\infty\) with pdf \(r(G(t))\) such that \(r(G(t))=-\log [1-G(t)]\) is monotonically non-decreasing; \(r(G(t))\) is closed in the interval \([m, n]\); and \(r(G(t))\) approaches \(m\) as \(t\) tends to negative infinity, and \(r(G(t))\) approaches \(n\) as \(t\) tends to positive infinity. Thus, by 4 the cdf and the pdf of the WGG generated class of distribution can be expressed as
\[
\begin{equation*}
F(t)=1-\left[1+\frac{\lambda\left[(1-G(t))^{-\beta}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]} t>0, \lambda, \beta>0 \tag{3}
\end{equation*}
\]
and
\[
\begin{equation*}
f(t)=\frac{\lambda^{2} \beta}{(1+\beta \lambda)(1-G(t))^{(1+\beta)}} g(t)\left((1-G(t))^{-\beta}+\beta-1\right) e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]} \tag{4}
\end{equation*}
\]
for \(t>0, \lambda, \beta>0\), where \(g(t)\), and \(G(t)\) are the parents pdf and cdf.
The WGG generated reliability model can be expressed as
\[
\begin{equation*}
S_{W G G}(t)=\left[1+\frac{\lambda\left[(1-G(t))^{-\beta}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]} t>0, \lambda, \beta>0 \tag{5}
\end{equation*}
\]

The hazard rate function that corresponds to the WGG generated model is defined as
\[
\begin{equation*}
h_{W G G}(t)=\frac{\frac{\lambda^{2} \beta}{(1+\beta \lambda)(1-G(t))^{(1+\beta)}} g(t)\left((1-G(t))^{-\beta}+\beta-1\right)}{\left[1+\frac{\lambda\left[(1-G(t))^{-\beta}-1\right]}{1+\beta \lambda}\right]} t>0, \lambda, \beta>0 . \tag{6}
\end{equation*}
\]

The reversed hazard rate function is obtained as
\[
\begin{equation*}
r_{W G G}(t)=\frac{\frac{\lambda^{2} \beta}{(1+\beta \lambda)(1-G(t))^{(1+\beta)}} g(t)\left((1-G(t))^{-\beta}+\beta-1\right) e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]}}{1-\left[1+\frac{\lambda\left[(1-G(t))^{-\beta}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]}}, \tag{7}
\end{equation*}
\]
for \(t>0, \lambda, \beta>0\).
The cumulative hazard rate function of the WGG generated function is give as:
\[
\begin{equation*}
H_{W G G}(t)=\log (1+\beta \lambda)-\log \left([1+\beta \lambda]+\lambda\left[(1-G(t))^{-\beta}-1\right]\right)+\lambda\left([1-G(t)]^{-\beta}-1\right) \tag{8}
\end{equation*}
\]

\section*{3. The Quantile Function}

Quantile is fundamental for the simulation and estimation of a distribution parameter(s). Hence, it is a function that associates the probability distribution function of the WGG generated model of a random variable \(T\) such that the probability of the variable being less than or equal to that value equals the probability for a uniform interval \(q \in(0,1)\) is defined as
\[
\begin{equation*}
t=G^{-1}\left[1-\left[\frac{W_{-1}\left((q-1)(1+\beta \lambda) e^{(1+\beta \lambda)}\right)-(1+\beta \lambda)}{\lambda}+1\right]^{-\frac{1}{\beta}}\right] \tag{9}
\end{equation*}
\]
where \(W_{-1}\) is the Lambert-W or omega function as defined in 13 and 16 such that \(W(t)=e^{(W(t))}=t \in[-1, \infty)\).

In particular, the median is obtained when \(q=0.5\).
Theorem 1. The shape, characteristics, and behaviour of the \(W G G\) generated model can be examined by investigating the first and second derivatives of the log of the \(W G G\) generated pdf model. Thus, for \(f^{\prime}(t)<0\). Then, then \(c d f F(t)\) will be decreasing monotonically for all values of \(t\). The \(W G G\) generated model will be bimodal if \(f^{\prime \prime}(t)\) changes its signs from negative to non-negative, viz-a-viz.

Proof. The \(\log f(t)\) is give as
\[
\begin{aligned}
\log f(t) & =2 \log \lambda+\log \beta-\log (1+\beta \lambda)+\log g(t)-(1+\beta \lambda) \log (1-G(t)) \\
& +\log \left([1-G(t)]^{-\beta}+\beta-1\right)-\lambda\left([1-G(t)]^{-\beta}-1\right)
\end{aligned}
\]

Thus, taking the derivative with respect to the variable, we have
\[
\frac{\partial \log f(t)}{\partial t}=\frac{g^{\prime}(t)}{g(t)}+(1+\beta \lambda) \frac{g(t)}{S(t)}+\frac{\beta g(t) S^{-\beta-1}(t)}{S^{\beta}(t)+\beta-1}-\lambda g(t) S^{-\beta-1}(t)
\]
where \(S(t)=1-G(t)\). Hence, \(f^{\prime}(t)<0\) if \(g(t)<0\).

The second derivative was implemented to determine if the model was bimodal. Thus, the second derivative is given as
\[
\begin{aligned}
\frac{\partial^{2} \log f(t)}{\partial t^{2}}= & \frac{g^{\prime \prime}(t)}{g(t)}-\frac{1}{g^{\prime}(t)}+(1+\beta \lambda)\left[\frac{g^{\prime}(t)}{S(t)}+\frac{g^{2}(t)}{S^{2}(t)}\right]-\frac{\lambda g^{\prime}(t)}{S^{(\beta+1)}(t)} \\
& +\lambda(1+\beta \lambda) \frac{g^{2}(t)}{S^{(\beta+2)}(t)}+\frac{\beta g^{\prime}(t) S^{-\beta-1}(t)}{S^{\beta}(t)+\beta-1} \\
+ & \beta(\beta+1) \frac{g^{2}(t) S^{-(\beta+2)}(t)}{S^{\beta}(t)+\beta-1}+\frac{\beta^{2} g^{2}(t) S^{-2(\beta+1)}(t)}{\left(S^{\beta}(t)+\beta-1\right)^{2}}
\end{aligned}
\]

\section*{4. Order statistics}

Order statistics are useful tools to improve the robustness of sampling plans by variables, and shorten test times of Poisson processes.

Let \(T_{(1)}, T_{(2)}, T_{(3)}, \ldots, T_{(k)}\) be the order statistics for a random variable \(T_{1}, T_{2}, T_{3}, \ldots, T_{k}\) with WGG distribution. Then, the WGG density of the \(u^{t h}\) order statistics is given as
\[
\begin{equation*}
f_{u}(t)=\frac{k!}{(u-1)!(k-u)!} F^{u-1}(t) S^{k-u}(t) f(t)-\infty<t<\infty \tag{10}
\end{equation*}
\]

However, using the binomial expansion, and noting that \(S=1-G(t)\), we have the order statistics as
\[
\begin{align*}
f_{u}(t) & =\frac{\beta \lambda^{2} S^{-(\beta+1)} k!}{(u-1)!(k-u)!}\left(S^{-\beta}+\beta-1\right) \sum_{j=0}^{u-1}(-1)^{u-j-1}\binom{u-1}{j}  \tag{11}\\
& \times\left[(1+\beta \lambda)+\lambda\left(S^{-\beta}-1\right)\right]^{k+j-u+1} e^{-\lambda\left(S^{-\beta}-1\right)(k+j-u+1)}
\end{align*}
\]

The minimum order statistics is obtained when \(u=1\), and the maximum order statistics is obtained when \(u=k\) respectively.
4.1. Record value distributions of the WWG model. Let \(T_{i}\) for \(i=1,2,3, \ldots, k\) be a finite sequence of independently identically distributed random variables with WGG generated cdf \(F(t)\) and a record times given as \(U(1)=1\) and \(U(k+1)=\) \(\min \left\{j>U(k) ; T_{j}>T_{u(k)}\right\} ; k \in \mathbb{N}\) with the random variable \(T_{u(k)}(k \in \mathbb{N})\) as the upper record values. Then, the pdf of the \(i\) upper record value \(U R_{i}=T_{u(k)}\) with a
special case of \(U R_{1}=T_{1}\) is given as
\[
\begin{align*}
f_{U R_{i}}(t) & =\frac{f(t)}{\Gamma(i)}\{-\log [1-F(t)]\}^{i-1} \\
& =\frac{\lambda^{2} \beta g(t)\left((1-G(t))^{-\beta}+\beta-1\right) e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]}}{(1+\beta \lambda)(1-G(t))^{(1+\beta)} \Gamma(i)}  \tag{12}\\
& \times\left\{-\log \left[\left[1+\frac{\lambda\left[(1-G(t))^{-\beta}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[(1-G(t))^{-\beta}-1\right]}\right]\right\}^{i-1}
\end{align*}
\]

\section*{5. Sub-models}

Some special sub-models were considered for flexibility, viability, and tractability using the proposed WGG generated model. We present some special cases of the WGG generated family of distributions since it extends several useful distributions in the literature. For all cases listed next, we consider \(t, \lambda, \beta>0\). Especially submodels with increasing, decreasing shaped data with or without a flat region in modeling. These special sub-models include Burr-XII, Lomax, and Frechet distributions.
5.1. Weighted Gompertz-G Burr-XII (WGG-B) distribution. Consider the Burr XII distribution with positive parameters \(\theta\) and \(\rho\), and cdf and pdf given as \(G(t)=1-\left(1+t^{\theta}\right)^{-\rho}\) and \(g(t)=\theta \rho t^{\theta-1}\left(1+t^{\theta}\right)^{-\rho-1}\). Then, inserting these expressions into Equations (3) and (4) gives the WGG-B density function with the cdf and pdf given as
\[
\begin{equation*}
F(t)=1-\left[1+\frac{\lambda\left[\left(1+t^{\theta}\right)^{\beta \rho}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[\left(1+t^{\theta}\right)^{\beta \rho}-1\right]}, t>0, \lambda, \beta, \theta, \rho>0 \tag{13}
\end{equation*}
\]
and
\[
\begin{align*}
f(t) & =\frac{\lambda^{2} \beta\left(1+t^{\theta}\right)^{\rho(1+\beta)}}{(1+\beta \lambda)}\left(\left(1+t^{\theta}\right)^{\beta \rho}+\beta-1\right) e^{-\lambda\left[\left(1+t^{\theta}\right)^{\beta \rho}-1\right]}  \tag{14}\\
& \times \theta \rho t^{\theta-1}\left(1+t^{\theta}\right)^{-\rho-1}, t>0, \lambda, \beta, \theta, \rho>0
\end{align*}
\]

Plots of the WGG-B density function for the selected parameter values are displayed in Figure 1a. Figure 1b displays the corresponding hazard rate function (hrfs) for particular values of the parameters. The shapes of the hazard rate function indicated increasing, and decreasing.
5.2. Weighted Gompertz-G Lomax (WGG-L) distribution. Consider the Lomax distribution with positive shape parameters \(\theta\) and scale parameter \(\rho\), and cdf and pdf given as \(G(t)=1-\left(1+\frac{t}{\rho}\right)^{-\rho}\) and \(g(t)=\frac{\theta}{\rho}\left[1+\frac{t}{\rho}\right]^{-(\theta+1)}\). Then, inserting these expressions into Equations (3) and (4) gives the WGG-L density function with the cdf and pdf given as


Figure 1. The plots WGG-B model for selected values of parameters
\[
\begin{equation*}
F(t)=1-\left[1+\frac{\lambda\left[\left(1+\frac{t}{\rho}\right)^{\beta \rho}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[\left(1+\frac{t}{\rho}\right)^{\beta \rho}-1\right]}, t>0, \quad \lambda, \beta, \theta, \rho>0 \tag{15}
\end{equation*}
\]
and
\[
\begin{align*}
f(t) & =\frac{\lambda^{2} \beta\left(1+\frac{t}{\rho}\right)^{\rho(1+\beta)}}{(1+\beta \lambda)}\left(\left(1+\frac{t}{\rho}\right)^{\beta \rho}+\beta-1\right) e^{-\lambda\left[\left(1+\frac{t}{\rho}\right)^{\beta \rho}-1\right]}  \tag{16}\\
& \times \frac{\theta}{\rho}\left[1+\frac{t}{\rho}\right]^{-(\theta+1)}, t>0, \lambda, \beta, \theta, \rho>0
\end{align*}
\]

Plots of the WGG-L density function for the selected parameter values are displayed in Figure 2a. Figure 2b displays the corresponding hrfs for some particular values of the parameters. The shapes of the hazard rate function indicated increasing, and decreasing.
5.3. Weighted Gompertz-G Frechet (WGG-F) distribution. Consider the Frechet distribution with positive shape parameters \(\theta\) and scale parameter \(\rho\), and cdf and pdf given as \(G(t)=e^{-\left(\frac{\theta}{t}\right)^{\rho}}\) and \(g(t)=\rho \theta^{\rho} t^{-\rho-1} e^{-\left(\frac{\theta}{t}\right)^{\rho}}\). Then, inserting these expressions into Equations (3) and (4) gives the WGG-F density function with the cdf and pdf given as


Figure 2. The plots WGG-L model for selected values of parameters
\[
\begin{equation*}
F(t)=1-\left[1+\frac{\lambda\left[\left(1-e^{-\left(\frac{\theta}{t}\right)^{\rho}}\right)^{-\beta}-1\right]}{1+\beta \lambda}\right] e^{-\lambda\left[\left(1-e^{-\left(\frac{\theta}{t}\right)^{\rho}}\right)^{-\beta}-1\right]}, t>0, \lambda, \beta, \theta, \rho>0 \tag{17}
\end{equation*}
\]
and
\[
\begin{align*}
f(t) & =\frac{\lambda^{2} \beta}{(1+\beta \lambda)\left(1-e^{-\left(\frac{\theta}{t}\right)^{\rho}}\right)^{(1+\beta)}}\left(\left(1-e^{-\left(\frac{\theta}{t}\right)^{\rho}}\right)^{-\beta}+\beta-1\right) e^{-\lambda\left[\left(1-e^{-\left(\frac{\theta}{t}\right)^{\rho}}\right)^{-\beta}-1\right]} \\
& \times \rho \theta^{\rho} t^{-\rho-1} e^{-\left(\frac{\theta}{t}\right)^{\rho}}, t>0, \lambda, \beta, \theta, \rho>0 \tag{18}
\end{align*}
\]

Plots of the WGG-F density function for the selected parameter values are displayed in Figure 3a. Figure 3b displays the corresponding hrfs for some particular values of the parameters. The shapes of the hazard rate function indicated an increase.

\section*{6. Mathematical Expression}

To examine the productivity of the WGG generated model, mathematical expansion of the pdf and cdf is carried out. The exponential term in (3) and (4) can


Figure 3. The plots WGG-F model for selected values of parameters
be expressed as
\[
e^{-\lambda\left((1-G(t))^{-\beta}-1\right)}=\sum_{w=0}^{\infty} \frac{(-1)^{w} \lambda^{w}\left((1-G(t))^{-\beta}-1\right)^{w}}{w!}
\]

Also, by binomial expansion, we have
\[
\left((1-G(t))^{-\beta}-1\right)^{w}=\sum_{u=0}^{w}(-1)^{w-u}\binom{w}{u}(1-G(t))^{-u \beta}
\]

Hence, the WGG generated pdf can be expressed as power function as
\[
\begin{equation*}
f(t)=\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \alpha \mu_{(i, w, u)} g(t) G^{i}(t) \tag{19}
\end{equation*}
\]
where
\[
\alpha=\frac{\Gamma(u \beta+2 \beta+i+1)}{\Gamma(u \beta+2 \beta+1)}+(\beta-1) \frac{\Gamma(u \beta+\beta+i+1)}{\Gamma(u \beta+\beta+1)}
\]
and
\[
\mu_{(i, w, u)}=(-1)^{2 w-u+i}\binom{w}{u} \frac{\lambda^{w+2}}{i!w!} \frac{\beta}{(1+\beta \lambda)}
\]
where \(\Gamma(\cdot)\) is a gamma function.

\section*{7. Statistical Properties}

The viability and performance of the proposed model will be investigated by examining some general statistical properties of the WGG generated model in this section.

Oftentimes, the expectation, variance, and moments of random variables can be obtained from some characteristics of the distribution function. Some of these functions are the probability generating function and the moment generating function.

Lin's condition random variable The Lin's function for a pdf \(f\) of a random variable \(T\) with a support \(t>0\) is defined as
\[
L_{f(t)}=-t \frac{f^{\prime}(t)}{f(t)}=-t \sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \alpha \mu_{(i, w, u)} \frac{i g^{2}(t) G^{i-1}(t)+g^{\prime}(t) G^{i}(t)}{g(t) G^{i}(t)}
\]

Incomplete moments
The incomplete moments of the WGG generated model allow the shape of the moments of WGG generated distribution, which is of interest for many areas, including econometrics, finance, and reliability, to be visible.

The \(k^{t h}\) incomplete moment, say \(\tau_{k}(t)\) of the WGG generated moment is given as
\[
\tau_{k}(y)=\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \alpha \mu_{(i, w, u)} \eta_{k, i}(y)
\]
where \(\eta_{k, i}=\int_{0}^{y} t^{k} g(t) G^{i}(t) d t\).
Probability generating function
This is a useful mechanism for characterizing the distribution of the random variable \(T\) with the WGG generated model. It can succinctly be used to describe the sequence of the probability of the random variable \(T\) with the WGG distribution. Hence, a random variable \(T\) with a WGG distribution has the probability generating function defined as
\[
\begin{align*}
P(z) & =\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \int_{0}^{\infty} z^{t} \alpha \mu_{(i, w, u)} g(t) G^{i}(t) d t \\
& =\sum_{i, w, a=0}^{\infty} \sum_{u=0}^{w} \frac{(\log z)^{a} \alpha \mu_{(i, w, u)}}{a!} \int_{0}^{\infty} t^{a} g(t) G^{i}(t) d t  \tag{20}\\
& =\sum_{i, w, a=0}^{\infty} \sum_{u=0}^{w} \frac{(\log z)^{a} \alpha \mu_{(i, w, u)}}{a!} p(z)
\end{align*}
\]
where
\[
p(z)=\int_{0}^{\infty} t^{a} g(t) G^{i}(t) d t|z| \leq 1
\]

Moment generating function

The probability density function of the random variable \(T\) can be identified using the moment generating function instrument. This is, however, possible since the moment generating function is a non-negative integral of measurable function. Thus, for a random variable \(T\) with a WGG distribution, the moment generating function is given as
\[
\begin{align*}
M_{T}(z) & =\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \int_{0}^{\infty} e^{z t} \alpha \mu_{(i, w, u)} g(t) G^{i}(t) d t \\
& =\sum_{i, w, a=0}^{\infty} \sum_{u=0}^{w} \frac{z^{a} \alpha \mu_{(i, w, u)}}{a!} \int_{0}^{\infty} t^{a} g(t) G^{i}(t) d t  \tag{21}\\
& =\sum_{i, w, a=0}^{\infty} \sum_{u=0}^{w} \frac{z^{a} \alpha \mu_{(i, w, u)}}{a!} p(z)
\end{align*}
\]

Probability weighted moments
One of the widely used characteristics of a distribution is called L-moments or probability weighted moments. This characteristic is used in hydrology to estimate the parameters of flood distributions. This might be because it is less sensitive to outliers, lower sampling variability, and fast convergence to asymptotic normality. The shape of the WGG generated probability distribution can also be summarized using the L-moments. Thus, L-moments are defined as:
\[
\begin{equation*}
P_{w m}(w, v)=\int_{0}^{\infty} t^{w} F^{v}(t) f(t) d t \tag{22}
\end{equation*}
\]

However, \(F^{v}\) can be expressed as
\[
F^{v}=\sum_{i, w=0}^{\infty} \sum_{u=0}^{w+p} \sum_{p=0}^{v}(-1)^{2 w+v+i-u}\binom{v}{p}\binom{w+p}{u} \frac{\lambda^{w+p} p^{w} \Gamma(k \beta+i)}{w!i!\Gamma(k \beta)(1+\beta \lambda)^{p}} G^{i}(t)
\]
where \(\Gamma(\cdot)\) is a gamma function. Hence, L-moments is given as
\[
\begin{equation*}
P_{w m}(w, v)=\sum_{i, w=0}^{\infty} \sum_{u=0}^{w+p} \sum_{p=0}^{v} R_{(i, w, u, p)} \alpha \mu_{(i, w, u)} T_{i} \tag{23}
\end{equation*}
\]
where
\[
T_{i}=\int_{0}^{\infty} t^{w} g(t) G^{2 i}(t) d t
\]
and
\[
R_{(i, w, u, p)}=(-1)^{2 w+v+i-u}\binom{v}{p}\binom{w+p}{u} \frac{\lambda^{w+p} p^{w} \Gamma(k \beta+i)}{w!i!\Gamma(k \beta)(1+\beta \lambda)^{p}}
\]

\section*{Entropies}

The heterogeneity or impurity of the target variable of Poisson process can be measured by the amount of uncertainty associated in the value of a random variable.

Thus, the Shannon entropy of WGG generated random variable \(T\) is defined as
\[
\begin{equation*}
S_{e}(T)=E\left[-\sum_{i, w=0}^{\infty} \sum_{u=0}^{w}\left(\log \mu_{(i, w, u)}+\log \mu+\log g(t)+i \log G(t)\right)\right] \tag{24}
\end{equation*}
\]

The Renyi entropy is a measure that increasingly weighs all WGG generated random events with nonzero probability. As \(\theta\) approaches zero, the WGG generated Renyi entropy is given as
\[
\begin{equation*}
R_{\theta}=\frac{1}{(1-\theta)} \log \int_{0}^{\infty} f^{\theta}(t) d t \theta>0, \theta \neq 0 \tag{25}
\end{equation*}
\]

This implies
\[
\begin{align*}
R_{\theta} & =\frac{1}{(1-\theta)} \log \int_{0}^{\infty}\left(\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \mu_{(i, w, u)} \alpha g(t) G^{i}(t)\right)^{\theta} d t \\
& =\frac{1}{(1-\theta)} \log \left[\left(\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \mu_{(i, w, u)} \alpha\right)^{\theta} \int_{0}^{\infty} g(t)^{\theta} G^{i}(t)^{\theta}\right] d t  \tag{26}\\
& =\frac{1}{(1-\theta)} \log \left[\left(\sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \mu_{(i, w, u)} \alpha D_{i}\right)^{\theta}\right]
\end{align*}
\]
where
\[
D_{i}=\int_{0}^{\infty} g(t) G^{i}(t) d t . i=1,2,3, \ldots
\]

Moment of the residual In reliability theory, and life testing scenarios, the additional lifetime a process or a product that a component or chain has survived up to time \(t\) is called the vitality function or residual life function or truncated moment. It can also be used to obtain the distribution function \(F(t)\). Thus, the \(k^{t h}\) moment of the residual life defined as \(M_{r s}(x)=E\left[(T-x)^{k} \mid T \geq t\right]\). Hence, it is expressed as
\[
\begin{align*}
M_{r s}(x) & =\frac{1}{1-F(x)} \int_{x}^{\infty}(T-x)^{k} f(t) d t=\frac{1}{1-F(x)} \sum_{a=1}^{k}(-1)^{k-a} x^{k-a} \int_{x}^{\infty} t^{a} f(t) d t \\
& =\frac{\alpha}{1-F(x)} \sum_{i, w=0}^{\infty} \sum_{u=0}^{w} \sum_{a=1}^{k}(-1)^{k-a} x^{k-a} \mu_{(i, w, u)} \Im_{i} \tag{27}
\end{align*}
\]
where
\[
\Im_{i}=\int_{x}^{\infty} t^{a} g(t) G^{i}(t) d t
\]

Theorem 2. Let \(T\) be a random variable with a \(W G G\) generated probability distribution function \(F(t)\). Let \(S(t)=1-F(t)\) and \(M_{k}(y)=E\left[(T-y)^{k} \mid T>y\right], y \geq 0\).

Then,
\[
\frac{M_{k}^{\prime}(y)+k M_{k}(y)}{M_{k}(y)}=\frac{M_{k-1}^{\prime}(y)+(k-1) M_{k-1}(y)}{M_{k-1}(y)} \text { or }
\]
equivalently,
\[
M_{k-1}^{\prime}(y)=-(k-1) M_{k-2}^{\prime}(y)+\frac{M_{k}^{\prime}(y)}{M_{k}(y)} M_{k-1}(y)+\frac{k M_{k-1}^{2}(y)}{M_{k}(y)}
\]

Proof. Let
\[
M_{k}(y)=\frac{1}{S(y)} \int_{y}^{\infty} k(t-y)^{k-1} S(t) d t
\]

Then,
\[
\log M_{k}(y)=\log \int_{y}^{\infty} k(t-y)^{k-1} S(t) d t-\log S(y)
\]

Thus, differentiating with respect to \(y\), we have
\[
\frac{M_{k}^{\prime}(y)}{M_{k}(y)}=\frac{\int_{y}^{\infty}-k(k-1)(t-y)^{k-2} S(t) d t}{\int_{y}^{\infty} k(t-y)^{k-1} S(t) d t}-\frac{S^{\prime}(y)}{S(y)}=\frac{-k M_{k-1}(y)}{M_{k}(y)}-\frac{S^{\prime}(y)}{S(y)}
\]

Hence,
\[
\frac{M_{k}^{\prime}(y)+k M_{k-1}(y)}{M_{k}(y)}=-\frac{S^{\prime}(y)}{S(y)}=\frac{M_{k-1}^{\prime}(y)+(k-2) M_{k-2}(y)}{M_{k-1}(y)}
\]

\section*{8. Parameter Estimation}

It is intuitive to note that the parameters of the WGG generated model are descriptive measures of the entire population that determine the shape and location of the curve on the plot of the WGG generated distribution. Hence, for a better forecasting and regression analysis of the proposed WGG model to be efficient, there is a need to obtain the parameter estimates of the WGG generated model. Thus, in this section, the parameters of the WGG generated model are estimated using the maximum likelihood estimation (MLE) method.
8.1. Maximum Likelihood. Let \(\mathbf{T}=\left(T_{1}, T_{2}, \ldots, T_{k}\right)\) be a random sample obtained from the WGG generated distribution with unknown parameter vector \(\Theta=\) \((\beta, \lambda, \psi)^{T}\). Let \(\mathbf{t}=\left(t_{1}, t_{2}, \ldots, t_{k}\right)\) be a sample value of a random sample \(\mathbf{T}\). Then,
we can obtain the log-likelihood as
\[
\begin{align*}
\ell= & 2 k \log \lambda+k \log \beta+\sum_{a=1}^{k} \log g\left(t_{a}, \psi\right)-k \log (1+\beta \lambda) \\
& -(1+\beta \lambda) \sum_{a=1}^{k} \log \left(1-G\left(t_{a}, \psi\right)\right)+\sum_{a=1}^{k} \log \left(\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta}+\beta-1\right)  \tag{28}\\
& -\sum_{a=1}^{k} \lambda\left(\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta}-1\right) .
\end{align*}
\]

The parameters of the WGG generated model are obtained by taking the first partial derivative of the log-likelihood of the WGG model with respect to each of the parameters and equate to zero. Thus, we have
\[
\begin{align*}
\frac{\partial \ell}{\partial \lambda}= & \frac{2 k}{\lambda}-\frac{k \beta}{1+\beta \lambda}-\beta \sum_{a=1}^{k} \log \left(1-G\left(t_{a}, \psi\right)\right)-\sum_{a=1}^{k}\left(\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta}-1\right)=0  \tag{29}\\
\frac{\partial \ell}{\partial \psi}= & \sum_{a=1}^{k} \frac{g^{\prime}\left(t_{a}, \psi\right)}{g\left(t_{a}, \psi\right)}+\beta \sum_{a=1}^{k} \frac{g\left(t_{a}, \psi\right)\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta-1}}{\left(\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta}+\beta-1\right)}  \tag{30}\\
& +(1+\beta \lambda) \sum_{a=1}^{k} \frac{g\left(t_{a}, \psi\right)}{1-G\left(t_{a}, \psi\right)}-\beta \lambda \sum_{a=1}^{k} g\left(t_{a}, \psi\right)\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta-1}=0
\end{align*}
\]
and
\[
\begin{align*}
\frac{\partial \ell}{\partial \beta}= & \frac{k}{\beta}-\lambda \sum_{a=1}^{k} \log \left(1-G\left(t_{a}, \psi\right)\right)+\sum_{a=1}^{k} \frac{\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta} \log \left(1-G\left(t_{a}, \psi\right)\right)}{\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta}+\beta-1}  \tag{31}\\
& -\frac{k \lambda}{1+\beta \lambda}-\lambda \sum_{a=1}^{k}\left(1-G\left(t_{a}, \psi\right)\right)^{-\beta} \log \left(1-G\left(t_{a}, \psi\right)\right)=0
\end{align*}
\]

However, the solutions to the nonlinear equations (29), (30), and (31) are obtained in closed form using numerical methods. These numerical methods are beyond the scope of this article.

\section*{9. Applications}

The viability, tractability, and performance of the WGG generated model is examined by first performing a Monte Carlo simulation of some sub-models of the proposed model. The real-life applications of some of the sub-models of the proposed model were investigated and compared to some competitive-related models in the literature. The WGG sub-models were compared with some existing models based on their mean squared errors in the simulation cases and goodness-of-fit test statistics in life applications.
9.1. Simulation study. A Monte Carlo simulation was carried out to test the flexibility and efficiency of the proposed distribution. The simulation was achieved using the quantile function in (9) to generate random data for the proposed model with \(0<q<1\) for various values of \(\lambda=1.0, \beta=1.0, \theta=0.2\) and \(\rho=1.0\) for the Burr XII sub-model. \(\lambda=0.9, \beta=2.3, \theta=0.1\) and \(\rho=0.01\) for the Lomax sub-model, and \(\lambda=0.1, \beta=0.1, \theta=0.3\) and \(\rho=0.8\) for the Frechet sub-model for 1000 replicated trials.

The sample size \(n\) are taken as \(n=5,10,20,50,100,150,200,250,300,350,400,450\), and 500 The simulation studied the mean estimated (ME), biases, and mean squared errors (MSE). The result of the simulation is as shown in Table 1. In Table one, we observed that the biases converge to zero as sample sizes increase. The estimated mean also converges to the true value as the sample sizes increases. The mean square errors converge to zero.

The bias is obtained for \((W=\lambda, \beta, \theta, \rho)\) as
\[
\operatorname{Bia}_{W}=\frac{1}{1000} \sum_{i=1}^{1000}\left(\hat{W}_{i}-W\right)
\]

Also, the MSE is obtained as
\[
M \hat{S} E_{W}=\frac{1}{1000} \sum_{i=1}^{1000}\left(\hat{W}_{i}-W\right)^{2}
\]
9.2. Life applications. In most cases of statistical modeling, the interest is to estimate the model parameters and evaluate their test statistics goodness-of-fit. Thus, in this section, the viability, tractability, and effectiveness of the proposed model is investigated with the illustration of real-life data sets. The measures of the test statistics' goodness-of-fit were examined with some existing neighbourhood models in the literature. These models in the literature include, but are not limited to, the class of Weibull, Gompertz, Kumaraswamy, and Frechet distributions. The test statistics considered include the Akaike information criterion (AIC), AndersonDarling (A), Cramer-von Mises (W), Kolmogrov-Smirnov (KS), and p-value (p-val). The larger the p-value and the smaller the test statistics the better the model fits the data.
9.2.1. Obesity Data. The first data consist of 22 obesity among children and adolescents aged 12-19 by selected characteristics: United States, selected between 2015 - 2018 as reported by 9 . The data are available in https://www.cdc.gov/nchs/hus /contents.htm-Table-027. The data were measured based on height and weight. The data are as follows:
18.9,15.1,23.1,9.8,25.7,26.9,19.8,16.0,19.2,12.0,28.0,29.2,
17.9,14.2,27.0,7.4,23.3,24.6,23.9,21.7,18.4,10.6.

The descriptive statistics of the data are given in Table 2.

Table 1. The mean estimates (ME), biases and mean squared errors (MSE) for \(\lambda, \beta, \theta\) and \(\rho\) with WGG generated sub-models
\begin{tabular}{|c|c|c|c|c|c|}
\hline Distribution & Parameters & n & ME & Bias & MSE \\
\hline \multirow[t]{13}{*}{Burr XII} & & 05 & 1.06020 .82880 .17621 .0033 & 0.36030 .32880 .83060 .7033 & 0.35970 .18110 .69070 .6271 \\
\hline & & 10 & 1.13400 .81150 .17180 .9558 & 0.35920 .31570 .22860 .6779 & 0.35440 .15810 .68730 .5358 \\
\hline & & 20 & 1.19510 .77080 .16940 .9466 & 0.35920 .31570 .12850 .6740 & 0.34740 .13030 .18710 .5326 \\
\hline & \(\lambda=1.0\) & 50 & 1.16010 .74160 .17150 .9556 & 0.22840 .11150 .12830 .0667 & 0.33490 .12550 .03700 .5226 \\
\hline & & 100 & 1.06150 .75580 .17240 .9740 & 0.12440 .01920 .12830 .0648 & 0.13450 .02400 .02690 .4176 \\
\hline & \(\beta=1.0\) & 150 & 1.02190 .77960 .17140 .9648 & 0.11590 .01780 .12820 .0582 & 0.04000 .02210 .02690 .3150 \\
\hline & & 200 & 0.97160 .79610 .17170 .9779 & 0.10890 .01300 .02780 .0575 & 0.03310 .02110 .02610 .2107 \\
\hline & \(\theta=0.2\) & 250 & 0.97560 .78930 .17290 .9582 & 0.01810 .01010 .01760 .0418 & 0.03170 .01800 .02560 .1012 \\
\hline & & 300 & 0.93970 .80780 .17220 .9667 & 0.00280 .00980 .01730 .0406 & 0.02590 .01640 .01520 .0196 \\
\hline & \(\rho=1.0\) & 350 & 0.94080 .81570 .17170 .9575 & 0.00210 .00960 .01010 .0466 & 0.01580 .01490 .01480 .0190 \\
\hline & & 400 & 0.94080 .81570 .17170 .9575 & 0.00100 .00780 .00620 .0298 & 0.00280 .01420 .01340 .0106 \\
\hline & & 450 & 0.98410 .80920 .17270 .9269 & 0.00010 .00580 .00580 .0269 & 0.00090 .00840 .01270 .0094 \\
\hline & & 500 & 0.99110 .80300 .17420 .9152 & 0.00010 .00160 .00380 .0152 & 0.00060 .00630 .01010 .0092 \\
\hline \multirow[t]{13}{*}{Lomax} & & 05 & 1.00552 .10540 .09560 .0143 & 0.70550 .47351 .30440 .0918 & 0.77140 .47651 .30500 .0919 \\
\hline & & 10 & 0.96312 .18520 .09590 .0109 & 0.66310 .47241 .30410 .0917 & 0.70780 .47401 .30460 .0917 \\
\hline & & 20 & 0.90852 .24290 .09630 .0098 & 0.60850 .47181 .30370 .0916 & 0.62870 .47271 .30420 .0917 \\
\hline & \(\lambda=0.9\) & 50 & 0.88012 .27350 .09510 .0084 & 0.60710 .10101 .30490 .0416 & 0.10910 .17200 .10430 .0216 \\
\hline & & 100 & 0.88062 .27240 .09580 .0082 & 0.02320 .02080 .01420 .0391 & 0.02540 .04200 .01380 .0116 \\
\hline & \(\beta=2.3\) & 150 & 0.88702 .27080 .09770 .0083 & 0.01960 .01920 .01230 .0313 & 0.02380 .03000 .01250 .0113 \\
\hline & & 200 & 0.88702 .27180 .09910 .0084 & 0.01330 .01840 .01090 .0212 & 0.01470 .02920 .01140 .0112 \\
\hline & \(\theta=0.1\) & 250 & 0.88962 .27100 .09870 .0084 & 0.01320 .01810 .00930 .0211 & 0.01360 .01980 .01110 .0111 \\
\hline & & 300 & 0.89622 .26920 .10100 .0087 & 0.00960 .01730 .00820 .0210 & 0.01220 .01800 .00920 .0110 \\
\hline & \(\rho=0.01\) & 350 & 0.89932 .26840 .10200 .0088 & 0.00700 .01590 .00800 .0199 & 0.01000 .01660 .00820 .0109 \\
\hline & & 400 & 0.90162 .26810 .10280 .0089 & 0.00300 .01290 .00720 .0195 & 0.00970 .01210 .00730 .0098 \\
\hline & & 450 & 0.90022 .26730 .10140 .0090 & 0.00060 .00570 .00660 .0191 & 0.00550 .00890 .00670 .0094 \\
\hline & & 500 & 0.90012 .26590 .10050 .0091 & 0.00010 .00540 .00550 .0157 & 0.00380 .00750 .00560 .0069 \\
\hline \multirow[t]{13}{*}{Frechet} & & 05 & 0.04880 .34180 .04391 .0031 & 0.15150 .44180 .11930 .8031 & 0.19780 .63050 .12010 .9035 \\
\hline & & 10 & 0.04850 .32820 .04720 .9886 & 0.15120 .42820 .11910 .7886 & 0.18990 .62330 .11950 .8846 \\
\hline & & 20 & 0.05560 .27750 .04670 .9320 & 0.14440 .37750 .11850 .7320 & 0.17400 .57720 .11930 .8254 \\
\hline & \(\lambda=0.1\) & 50 & 0.08300 .19310 .04530 .8610 & 0.11700 .29310 .11830 .0610 & 0.13550 .25970 .11890 .1302 \\
\hline & & 100 & 0.09730 .13320 .03840 .8298 & 0.10270 .23320 .11750 .0435 & 0.02270 .02890 .01880 .0141 \\
\hline & \(\beta=0.1\) & 150 & 0.10200 .11290 .03520 .8287 & 0.09800 .21290 .11740 .0423 & 0.02050 .02690 .01820 .0113 \\
\hline & & 200 & 0.10540 .10540 .03340 .8234 & 0.09460 .20240 .11660 .0402 & 0.01910 .02250 .01790 .0104 \\
\hline & \(\theta=0.3\) & 250 & 0.10690 .10460 .03260 .8244 & 0.09350 .19760 .11480 .0368 & 0.01770 .01380 .01760 .0094 \\
\hline & & 300 & 0.10840 .10360 .03250 .8352 & 0.09160 .20360 .11160 .0352 & 0.01540 .01310 .01670 .0083 \\
\hline & \(\lambda=0.8\) & 350 & 0.11020 .10240 .03170 .8068 & 0.08980 .20240 .10610 .0298 & 0.01240 .01100 .01600 .0074 \\
\hline & & 400 & 0.10410 .10160 .03150 .8012 & 0.08890 .20360 .10470 .0287 & 0.01130 .01020 .01530 .0066 \\
\hline & & 450 & 0.10210 .10140 .03090 .8003 & 0.08790 .20370 .10330 .0244 & 0.01000 .00990 .01370 .0053 \\
\hline & & 500 & 0.10060 .10070 .03070 .8005 & 0.08740 .20340 .10280 .0234 & 0.00920 .00910 .01270 .0029 \\
\hline
\end{tabular}

Table 2. The Descriptive statistics of obesity among children and adolescents data set to 2 decimal points
\begin{tabular}{cccccccccc}
\hline \hline Mean & Median & \(\sigma\) & IQR & Variance & Kurtosis & Skewness & \(25 \%\) & \(75 \%\) & \(99 \%\) \\
\hline 19.67 & 19.50 & 6.30 & 9.10 & 39.66 & -1.12 & -0.29 & 15.33 & 24.43 & 28.95 \\
\hline
\end{tabular}

We observed from Table 2 that the a negative kurtosis and skewness were obtained. This implies that the distribution of the obesity data is flatter than a normal curve with the same mean and standard deviation. Hence, the data are left skewed.

Table 3 shows the test statistics of the goodness-of-fit measure of comparison adopted for comprehensive comparison.

Table 3: The goodness-of-fit measure of obesity among children and adolescents data set (standard errors in parentheses)
\begin{tabular}{|c|c|c|c|c|c|c|}
\hline Distribution & p-value & AIC & KS & W & A & Estimates \\
\hline \multirow[t]{4}{*}{WGG-B} & \multirow[t]{4}{*}{0.9997} & \multirow[t]{4}{*}{97.9173} & \multirow[t]{4}{*}{0.0219} & \multirow[t]{4}{*}{0.0012} & \multirow[t]{4}{*}{0.0528} & \(\hat{\lambda}=0.2201(0.1022)\) \\
\hline & & & & & & \(\hat{\beta}=1.2315(0.0898)\) \\
\hline & & & & & & \(\hat{\rho}=0.0872(0.0098)\) \\
\hline & & & & & & \(\hat{\theta}=0.2124(0.0252)\) \\
\hline \multirow[t]{4}{*}{WGG-L} & \multirow[t]{4}{*}{0.9390} & \multirow[t]{4}{*}{102.3145} & \multirow[t]{4}{*}{0.1071} & \multirow[t]{4}{*}{0.0278} & \multirow[t]{4}{*}{0.2021} & \(\hat{\lambda}=0.0075(0.0020)\) \\
\hline & & & & & & \(\hat{\beta}=1.2912(0.0125)\) \\
\hline & & & & & & \(\hat{\rho}=0.1142(0.0967)\) \\
\hline & & & & & & \(\hat{\theta}=0.8494(0.3254)\) \\
\hline \multirow[t]{4}{*}{WGG-F} & \multirow[t]{4}{*}{0.9324} & \multirow[t]{4}{*}{109.8906} & \multirow[t]{4}{*}{0.2095} & \multirow[t]{4}{*}{0.0452} & \multirow[t]{4}{*}{0.3183} & \(\hat{\lambda}=0.0022(0.0004)\) \\
\hline & & & & & & \(\hat{\beta}=2.4857(0.8351)\) \\
\hline & & & & & & \(\hat{\rho}=0.8792(0.2743)\) \\
\hline & & & & & & \(\hat{\theta}=1.0516(0.2778)\) \\
\hline \multirow[t]{4}{*}{KB} & \multirow[t]{4}{*}{0.7640} & \multirow[t]{4}{*}{153.2259} & \multirow[t]{4}{*}{0.1356} & \multirow[t]{4}{*}{0.0811} & \multirow[t]{4}{*}{0.5349} & \(\hat{\alpha}=33.4661(17.9125)\) \\
\hline & & & & & & \[
\hat{\beta}=47.4488(46.2083)
\] \\
\hline & & & & & & \[
\hat{\rho}=0.0331(1.8429)
\] \\
\hline & & & & & & \(\hat{\theta}=21.8947(7.1942)\) \\
\hline \multirow[t]{4}{*}{KL} & \multirow[t]{4}{*}{0.6961} & \multirow[t]{4}{*}{154.7281} & \multirow[t]{4}{*}{0.1443} & \multirow[t]{4}{*}{0.1006} & \multirow[t]{4}{*}{0.6475} & \(\hat{\alpha}=14.5201(14.8943)\) \\
\hline & & & & & & \(\hat{\beta}=1.3267(1.7158)\) \\
\hline & & & & & & \(\hat{\rho}=0.0079(0.0030)\) \\
\hline & & & & & & \(\hat{\theta}=20.3753(16.9103)\) \\
\hline \multirow[t]{4}{*}{KF} & \multirow[t]{4}{*}{0.7788} & \multirow[t]{4}{*}{152.0259} & \multirow[t]{4}{*}{0.1336} & \multirow[t]{4}{*}{0.0642} & \multirow[t]{4}{*}{0.4347} & \(\hat{\alpha}=5.1639(6.2917)\) \\
\hline & & & & & & \(\hat{\beta}=166.4803(246.7566)\) \\
\hline & & & & & & \(\hat{\rho}=0.6187(0.1771)\) \\
\hline & & & & & & \(\hat{\theta}=20.9740(34.9274)\) \\
\hline \multirow[t]{4}{*}{KW} & \multirow[t]{4}{*}{0.6844} & \multirow[t]{4}{*}{147.3603} & \multirow[t]{4}{*}{0.0915} & \multirow[t]{4}{*}{0.0194} & \multirow[t]{4}{*}{0.1346} & \(\hat{\alpha}=0.2099(0.2886)\) \\
\hline & & & & & & \(\hat{\beta}=1.1818(1.3944)\) \\
\hline & & & & & & \(\hat{\rho}=0.0356(0.0084)\) \\
\hline & & & & & & \(\hat{\theta}=12.5159(17.5384)\) \\
\hline \multirow[t]{3}{*}{APG} & \multirow[t]{3}{*}{0.6866} & \multirow[t]{3}{*}{147.2333} & \multirow[t]{3}{*}{0.0901} & \multirow[t]{3}{*}{0.0296} & \multirow[t]{3}{*}{0.2029} & \(\hat{\alpha}=1.8905(2.7086)\) \\
\hline & & & & & & \(\hat{\beta}=0.0051(0.0037)\) \\
\hline & & & & & & \(\hat{\rho}=0.1627(0.0287)\) \\
\hline \multirow[t]{2}{*}{GB} & \multirow[t]{2}{*}{0.2511} & \multirow[t]{2}{*}{156.8906} & \multirow[t]{2}{*}{0.2095} & \multirow[t]{2}{*}{0.0452} & \multirow[t]{2}{*}{0.3183} & \(\hat{\alpha}=0.0022(0.0004)\) \\
\hline & & & & & & \(\hat{\beta}=2.4857(0.8351)\) \\
\hline
\end{tabular}

Table 3 - Continued from previous page


Figure 5 shows the empirical histogram and cdfs of the obesity real-life data applications.
9.2.2. Precipitations in Karachi city, Pakistan Data. The second data examined comprises 59 annual maximum precipitations in Karachi city, Pakistan, for the


Figure 4. The Empirical densities and cdfs of obesity among children and adolescents data set
years 1950-2009 as used in 6. The precipitation records help water management studies and flood defense systems to predict floods and droughts. The precipitation data also help to minimize the risk of large hydraulic structures. The values of the data are:
\(11.8,6.5,54.9,39.9,16.8,30.2,38.4,76.9,73.4,117.6,157.7,148.6,11.4,5.6\), \(63.6,62.4,85,256.3,24.9,148.6,160.5,131.3,77,155.2,217.2,105.5,166.8,157.9\), \(73.6,291.4\), , 30 , \(270.4,160,96.3,185.7,429.3,184.9,262.5,80.6,138.2,28,39.3\), \(210.3,315.7,107.7,33.3,302.6,159.1,78.7,33.2,52.2,92.7,150.4,43.7,68.3,20.8\), 179.4, 245.7, 19.5.

The descriptive statistics of the data are given in Table 4.
Table 4. The Descriptive statistics of annual maximum precipitations in Karachi city, Pakistan data set to 2 decimal points
\begin{tabular}{cccccccccc}
\hline \hline Mean & Median & \(\sigma\) & IQR & Variance & Kurtosis & Skewness & \(25 \%\) & \(75 \%\) & \(99 \%\) \\
\hline 118.40 & 92.70 & 93.21 & 120.65 & 8688.99 & 0.64 & 0.99 & 39.60 & 160.25 & 363.41 \\
\hline
\end{tabular}

We observed from Table 4 that the a positive kurtosis and skewness indicated that distribution is peaked and possesses thick tails, and most values are clustered around the left tail of the distribution while the right tail of the distribution is longer.

Table 5: The goodness-of-fit measure of maximum precipitations in Karachi city, Pakistan data set (standard errors in parentheses)
\begin{tabular}{|c|c|c|c|c|c|c|}
\hline Distribution & p-value & AIC & KS & W & A & Estimates \\
\hline \multirow[t]{4}{*}{WGG-B} & \multirow[t]{4}{*}{0.9470} & \multirow[t]{4}{*}{383.7050} & \multirow[t]{4}{*}{0.0961} & \multirow[t]{4}{*}{0.0454} & \multirow[t]{4}{*}{0.2577} & \(\hat{\lambda}=0.0951(0.1763)\) \\
\hline & & & & & & \(\hat{\beta}=1.2401(0.1875)\) \\
\hline & & & & & & \(\hat{\rho}=1.7186(0.2943)\) \\
\hline & & & & & & \(\hat{\theta}=1.9784(0.6677)\) \\
\hline \multirow[t]{4}{*}{WGG-L} & \multirow[t]{4}{*}{0.9376} & \multirow[t]{4}{*}{391.7440} & \multirow[t]{4}{*}{0.0968} & \multirow[t]{4}{*}{0.0461} & \multirow[t]{4}{*}{0.2616} & \(\hat{\lambda}=0.1689(0.0778)\) \\
\hline & & & & & & \(\hat{\beta}=1.2239(0.0682)\) \\
\hline & & & & & & \(\hat{\rho}=1.0199(0.0332)\) \\
\hline & & & & & & \(\hat{\theta}=1.5536(0.0941)\) \\
\hline \multirow[t]{4}{*}{WGG-F} & \multirow[t]{4}{*}{0.8617} & \multirow[t]{4}{*}{401.1031} & \multirow[t]{4}{*}{0.0989} & \multirow[t]{4}{*}{0.0688} & \multirow[t]{4}{*}{0.3093} & \(\hat{\lambda}=1.4858(0.4944)\) \\
\hline & & & & & & \(\hat{\beta}=1.2023(0.7353)\) \\
\hline & & & & & & \(\hat{\rho}=1.1434(0.1052)\) \\
\hline & & & & & & \(\hat{\theta}=1.3496(0.9372)\) \\
\hline \multirow[t]{4}{*}{KB} & \multirow[t]{4}{*}{0.2911} & \multirow[t]{4}{*}{691.8905} & \multirow[t]{4}{*}{0.1276} & \multirow[t]{4}{*}{0.1372} & \multirow[t]{4}{*}{0.8463} & \(\hat{\alpha}=8.3342(2.2157)\) \\
\hline & & & & & & \(\hat{\beta}=56.1819(92.7683)\) \\
\hline & & & & & & \(\hat{\rho}=0.0182(0.0000)\) \\
\hline & & & & & & \(\hat{\theta}=11.1408(1.0780)\) \\
\hline \multirow[t]{4}{*}{KL} & \multirow[t]{4}{*}{0.4207} & \multirow[t]{4}{*}{687.9069} & \multirow[t]{4}{*}{0.1145} & \multirow[t]{4}{*}{0.0848} & \multirow[t]{4}{*}{0.4997} & \(\hat{\alpha}=1.7166(0.2951)\) \\
\hline & & & & & & \(\hat{\beta}=3.3847(2.8572)\) \\
\hline & & & & & & \(\hat{\rho}=0.0040(0.0010)\) \\
\hline & & & & & & \(\hat{\theta}=1.5341(1.0421)\) \\
\hline \multirow[t]{4}{*}{KF} & \multirow[t]{4}{*}{0.3786} & \multirow[t]{4}{*}{687.6918} & \multirow[t]{4}{*}{0.1185} & \multirow[t]{4}{*}{0.0883} & \multirow[t]{4}{*}{0.5257} & \(\hat{\alpha}=6.8464(2.1692)\) \\
\hline & & & & & & \(\hat{\beta}=161.821(229.22)\) \\
\hline & & & & & & \(\hat{\rho}=0.2188(0.0564)\) \\
\hline & & & & & & \(\hat{\theta}=30.025(31.898)\) \\
\hline \multirow[t]{4}{*}{KW} & \multirow[t]{4}{*}{0.7467} & \multirow[t]{4}{*}{684.7171} & \multirow[t]{4}{*}{0.0883} & \multirow[t]{4}{*}{0.0467} & \multirow[t]{4}{*}{0.2692} & \(\hat{\alpha}=0.8755(0.4893)\) \\
\hline & & & & & & \(\hat{\beta}=0.5662(0.6176)\) \\
\hline & & & & & & \(\hat{\rho}=0.0112(0.0098)\) \\
\hline & & & & & & \(\hat{\theta}=1.3454(0.3905)\) \\
\hline \multirow[t]{3}{*}{APG} & \multirow[t]{3}{*}{0.8959} & \multirow[t]{3}{*}{682.9092} & \multirow[t]{3}{*}{0.0748} & \multirow[t]{3}{*}{0.0438} & \multirow[t]{3}{*}{0.2641} & \(\hat{\alpha}=1.5772(2.1911)\) \\
\hline & & & & & & \(\hat{\beta}=0.0073(0.0040)\) \\
\hline & & & & & & \(\hat{\rho}=0.0023(0.0022)\) \\
\hline \multirow[t]{4}{*}{GB} & \multirow[t]{4}{*}{0.6326} & \multirow[t]{4}{*}{684.8519} & \multirow[t]{4}{*}{0.0972} & \multirow[t]{4}{*}{0.0491} & 0.2803 & \(\hat{\alpha}=0.0075(0.0045)\) \\
\hline & & & & & & \(\hat{\beta}=2.7856(1.9958)\) \\
\hline & & & & & & \(\hat{\rho}=0.3543(0.3103)\) \\
\hline & & & & & & \(\hat{\theta}=1.2401(0.9676)\) \\
\hline
\end{tabular}

Table 5 - Continued from previous page
\begin{tabular}{|c|c|c|c|c|c|c|}
\hline Distribution & p -value & AIC & KS & W & A & Estimates \\
\hline \multirow[t]{4}{*}{GF} & 0.6774 & 683.6124 & 0.0937 & 0.0435 & 0.2460 & \[
\hat{\hat{\alpha}=0.1587(0.4962)}
\] \\
\hline & & & & & & \(\hat{\beta}=1.8235(2.5702)\) \\
\hline & & & & & & \(\hat{\rho}=0.7248(0.7488)\) \\
\hline & & & & & & \(\hat{\theta}=22.8034(61.9942)\) \\
\hline \multirow[t]{4}{*}{GL} & 0.7704 & 685.1274 & 0.0864 & 0.0488 & 0.2849 & \(\hat{\alpha}=0.1380(2.0119)\) \\
\hline & & & & & & \(\hat{\beta}=1.7962(38.1464)\) \\
\hline & & & & & & \(\hat{\rho}=0.0437(0.3829)\) \\
\hline & & & & & & \(\hat{\theta}=0.7748(16.8750)\) \\
\hline \multirow[t]{4}{*}{WF} & 0.7042 & 681.9136 & 0.0916 & 0.0434 & 0.2462 & \(\hat{\alpha}=0.0358(0.0180)\) \\
\hline & & & & & & \(\hat{\beta}=0.2947(0.1467)\) \\
\hline & & & & & & \(\hat{\rho}=4.1927(2.0935)\) \\
\hline & & & & & & \(\hat{\theta}=8.6209(1.7749)\) \\
\hline \multirow[t]{4}{*}{WB} & 0.5757 & 684.6917 & 0.1016 & 0.0519 & 0.2954 & \(\hat{\alpha}=0.0073(0.0049)\) \\
\hline & & & & & & \(\hat{\beta}=2.4377(1.0282)\) \\
\hline & & & & & & \(\hat{\rho}=0.4476(0.4885)\) \\
\hline & & & & & & \(\hat{\theta}=0.9901(1.3096)\) \\
\hline \multirow[t]{4}{*}{WL} & 0.1985 & 751.7122 & 0.1398 & 0.0730 & 0.4267 & \(\hat{\alpha}=3.6920(0.7601)\) \\
\hline & & & & & & \(\hat{\beta}=0.0923(0.0253)\) \\
\hline & & & & & & \(\hat{\rho}=0.6424(0.0600)\) \\
\hline & & & & & & \(\hat{\theta}=0.1421(0.5247)\) \\
\hline \multirow[t]{3}{*}{GE} & 0.3220 & 682.9042 & 0.0717 & 0.0420 & 0.2562 & \(\hat{\alpha}=0.0857(0.0178)\) \\
\hline & & & & & & \(\hat{\beta}=0.0438(0.0473)\) \\
\hline & & & & & & \(\hat{\rho}=0.0707(0.6964)\) \\
\hline \multirow[t]{4}{*}{GW} & 0.4824 & 687.6262 & 0.0604 & 0.0582 & 0.3764 & \(\hat{\alpha}=0.0341(0.0082)\) \\
\hline & & & & & & \(\hat{\beta}=0.0787(0.0160)\) \\
\hline & & & & & & \(\hat{\rho}=0.3342(0.0000)\) \\
\hline & & & & & & \(\hat{\theta}=0.7105(0.0072)\) \\
\hline \multirow[t]{4}{*}{TF} & 0.3617 & 701.1031 & 0.1201 & 0.2688 & 1.6393 & \(\hat{\alpha}=28.4858(29.4944)\) \\
\hline & & & & & & \(\hat{\beta}=31.2023(13.7353)\) \\
\hline & & & & & & \(\hat{\rho}=1.1434(0.1052)\) \\
\hline & & & & & & \(\hat{\theta}=0.9372(4.2815)\) \\
\hline
\end{tabular}

Figure 6 shows the empirical histogram and cdfs of the obesity real-life data applications.
9.3. Discussion. In Tables 3 and 5 , we observed that the p-values of the WGG generated models are the highest with the lowest AIC test statistic in Burr XII, Lomax, and Frechet sub-models. Hence, the WGG model has provided a better alternative to making statistical distributions more flexible, and viable compared


Figure 5. The Empirical densities and cdfs of maximum precipitations in Karachi city, Pakistan data set
to the model generated by Gompertz, Weibull, Kumaraswamy, and Alpha power models.

\section*{10. Conclusion}

Intuitively, a two-parameter weighted Gompertz-G generated distribution was examined and introduced by making use of a weighted Gompertz and the T-X characterizations. The newly developed model has found its uses in cases where two-sided abrupt changes schemes occurred in applications. The WGG model has provided a better alternative to making statistical distributions more flexible, and viable compared to the model generated by Gompertz, Weibull, Kumaraswamy, and Alpha power models. The statistical properties and estimations of the model parameters were obtained. The viability and flexibility of the WGG-generated model were demonstrated by illustration of a simulation and real data sets using their goodness-of-fit statistics. The outcomes of the WGG-generated test statistics indicated a better viable, tractable, flexible, and parsimonious generator compared to some competitive models in the literature. Hence, it can be used as a better alternative in reliability theory and extreme value theory.

Author Contribution Statements The authors contributed equally to this paper. All authors read and approved this paper's final form.

Declaration of Competing Interests The authors wish to state clearly that there is no conflict of interest.

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\title{
FRACTIONAL APPROACH FOR DIRAC OPERATOR INVOLVING M-TRUNCATED DERIVATIVE
}

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\begin{abstract}
In this study, we examine the basic spectral information for systems governed by the Dirac equation with distinct boundary conditions, utilizing a modified form of local derivatives known as M-truncated derivative (MTD). The spectral information discussed includes the representation of solutions in the form of integral equations, the asymptotics vector-valued eigenfunctions and eigenvalues, and their normalized forms, all within the context of the MTD method that incorporates truncated Mittag-Leffler functions. This type of MTD provides the features of integer-order operator theory. Also, by virtue of the parameters \(\alpha\) and \(\gamma\), we analyze and compare the solutions with graphs in terms of different potentials, different eigenvalues and different orders. Thus, the aim of this article is to consider spectral structure of Dirac system in frame of M -truncated derivative by proping with visual analysis.
\end{abstract}

\section*{1. Introduction}

Studies related to several types of differential equations are always attracted by scientists. Because the differential equations have the speciality to model more complex natural systems. Also, the main advantage of fractional derivatives is that it allows us to achieve better results in modeling. Many fractional integral and derivatives like Liouville-Caputo, Riemann-Liouville, Hilfer, Atangana-Baleanu, CaputoFabrizio, etc. has been introduced and studied by scientists in \(4,12,13,28\). In recently, Khalil et al. has described the local derivative, which is also referred to as the conformable derivative depending on the basic limit definitions of the derivative firstly in 20 . The conformable derivative is very useful in applied mathematics because it shows parallel features to the ordinary derivative like quotient of two functions and the derivative of the product. Also it enables changing of order between \(0<\alpha \leq 1\). Because of this reason, many scientists have applied this

\footnotetext{
2020 Mathematics Subject Classification. 34A08, 34A45, 34L30.
Keywords. M-truncated derivative, Dirac operator, spectral data, visual results.
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}
derivative to their studies like \(1-3,5,6,8,9,16,18,20\). Proportional \(\alpha\) - derivative has similar features with conformable derivative but it differs in its limit definition which presented by Katugampola 19 . It was studied in 6, 7. In recently, Mfractional derivative containing a Mittag-Leffler function with one parameter has been introduced by Sousa and Oliveira in \(25-27\). Benefiting from the definition of these four local derivatives as mentioned above, M-truncated derivative is introduced by Sousa and Oliveira in 25 and it represents a generalization of the other four local derivatives because of the additional parameter inside definition. All other definitions of local derivative such as Katugampola, M-truncated derivative are adaptations of conformable derivative. In these derivatives, basic formulas such as quotient of two function, derivative of the product, chain rule, Leibniz rule etc. shares similarities with conformable derivatives. Spectral analysis of M-truncated derivative for Sturm-Liouville problem and some applications containing truncated Mittag- Leffler function are studied in 24, 29 . 31.

Dirac equation has a big importance in the modern field of atomic physics. The deepest meaning of the Dirac equation was that any relative definition of a particle necessarily includes not only the wave function of a single particle, but also multiple wave functions representing the potential of other particles. Dirac equation systems have applications in many branches of science like electrical engineering, mathematics and physics. New applications of conclusions and opinions from this topic shed light on future problems such as inverse problems of spectral theory. A first-order matrix linear differential equation whose solution is a 4 -component wave function (a spinor) is so important in physics and mathematics 10, 14, 15.

Let \(L\) be a matrix operator defined by
\[
\left(\begin{array}{cc}
V(x)+m & 0 \\
0 & V(x)-m
\end{array}\right)
\]
where \(V(x)\) is a potential function, \(m\) is the mass of a particle and \(y(x)\) denote a two component vector function \(y(x)=\binom{y_{1}(x)}{y_{2}(x)}\). Then let's consider the equation
\[
\left(B \frac{d}{d x}+L-\lambda I\right) y=0
\]
where \(\lambda\) is a parameter and
\[
B=\left(\begin{array}{cc}
0 & 1 \\
-1 & 0
\end{array}\right), I=\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
\]
is equivalent to a system
\[
\begin{aligned}
\frac{d y_{2}(x, \lambda)}{d x}+(V(x)+m) y_{1}(x, \lambda) & =\lambda y_{1}(x, \lambda) \\
-\frac{d y_{1}(x, \lambda)}{d x}+(V(x)-m) y_{2}(x, \lambda) & =\lambda y_{2}(x, \lambda)
\end{aligned}
\]

The basic analysis of the spectral structure for the Dirac operator means that finding asymptotic behaviors of the eigenvalues, the vector-valued eigenfunctions and the norming constants and showing the reality of the eigenvalues and the ortogonality of the eigen-vector-functions, etc. This type of analysis is called a direct problem. In this article, the reality of the eigenvalues, and the orthogonality of the eigenfunctions have been shown and the asymptotic formulas for the eigenvalues, eigen-vector-functions, the normalized eigen-vector-functions and the norming constants have been obtained in terms of M-truncated derivative for Dirac system having separated boundary conditions. The studies on the direct and inverse eigenvalue problems can be viewed from \(11,21,23\). Basic spectral features of linear differential operators including conformable derivatives, which inspired our work, were studied by \(2,3: 5: 25\). Authors have established an existence and uniqueness theorem for a conformable fractional Dirac system in study 2. Also they have addresses the existence of a spectral function for a singular conformable Dirac system in 3. The M-truncated derivative can be employed in studies related to eigenvalue problems and spectral analysis. It is particularly beneficial in such analyses related to Dirac operators based on fractional derivatives. Our primary reason for selecting this particular local derivative is its inclusivity of other local derivatives, owing to the presence of an additional parameter associated with the Mittag-Leffler function. Differing from the literature, our results are more comprehensive compared to other local derivatives due to the presence of the parameter associated with the Mittag-Leffler function.

The layout of this research is presented in the following way: in section 2, we present to definitions and fundamental properties of MTD. In section 3, the spectral structure of Dirac system is studied. This main part of our study involves the reality of the eigenvalues, the orthogonality of the eigenvector-functions, and asymptotic formulas for the vector-valued eigen-functions, the eigenvalues, the norming constants and their normalized forms. Section 4 prerents detailed discussion about simulation anaysis by supporting with the graphs for different values of \(\alpha, \gamma\) and \(\lambda\). In part 5 , the remarks of main results close the paper.

\section*{2. Preliminaries}

In this part, we assign some necessary definitions, theorems and lemmas related to MTD.

Definition 1. 255 The concept of the truncated Mittag-Leffler function with a single parameter is introduced through,
\[
\begin{equation*}
{ }_{i} E_{\gamma}(z)=\sum_{k=0}^{i} \frac{z^{k}}{\Gamma(\gamma k+1)} . \tag{1}
\end{equation*}
\]

Definition 2. 25 Let \(f:[0, \infty) \rightarrow \mathbb{R}\) be a function for \(t>0\), then MTD of \(f\) with order \(0<\alpha \leq 1\) is defined by
\[
\begin{equation*}
{ }_{i} T_{M}^{\alpha, \gamma} f(t)=\lim _{\varepsilon \rightarrow 0} \frac{f\left(t_{i} E_{\gamma}\left(\varepsilon t^{-\alpha}\right)\right)-f(t)}{\varepsilon} \tag{2}
\end{equation*}
\]
where \({ }_{i} E_{\gamma}(\).\() is the truncated Mittag-Leffler function defined in (1) for \gamma>0\).
Definition 3. 25 The M-integral is defined as follows
\[
\left({ }_{M} I_{a}^{\alpha, \gamma} f\right)(\tau)=\int_{a}^{\tau} f(t) d_{\alpha, \gamma} t=\Gamma(\gamma+1) \int_{a}^{\tau} \frac{f(t)}{t^{1-\alpha}} d t
\]
where \(\gamma>0, \alpha \in(0,1]\), and \(f\) is defined in \((a, \tau]\).
Lemma 1. 255 Let \(\alpha \in(0,1], \gamma>0\) and \(f, g\) be \(\alpha\)-differentiable at a point \(t>0\).
Then,
1. \({ }_{i} T_{M}^{\alpha, \gamma}(a f+b g)=a_{i} T_{M}^{\alpha, \gamma} f+b_{i} T_{M}^{\alpha, \gamma}(g)\) for \(a, b \in \mathbb{R}\);
2. \({ }_{i} T_{M}^{\alpha, \gamma}\left(t^{n}\right)=n t^{n-\alpha}\) for all \(n \in \mathbb{R}\);
3. \({ }_{i} T_{M}^{\alpha, \gamma}(f g)=f_{i} T_{M}^{\alpha, \gamma}(g)+g_{i} T_{M}^{\alpha, \gamma}(f)\);
4. \({ }_{i} T_{M}^{\alpha, \gamma}\left(\frac{f}{g}\right)=\frac{g_{i} T_{M}^{\alpha, \gamma}(f)-f_{i} T_{M}^{\alpha, \gamma}(g)}{g^{2}}\);
5. \({ }_{i} T_{M}^{\alpha, \gamma}(c)=0, c\) is a constant;
6. \({ }_{i} T_{M}^{\alpha, \gamma}(f o g)(t)=f^{\prime}(g(t))_{i} T_{M}^{\alpha, \gamma} g(t)\), for \(f\) is differentiable at \(g(t)\);
7. If \(f\) is differentiable, thus \({ }_{i} T_{M}^{\alpha, \gamma}(f)(t)=\frac{t^{1-\alpha}}{\Gamma(\gamma+1)} \frac{d f(t)}{d t}\).

Theorem 1. Assume that \(f, g:[a, b] \rightarrow \mathbb{R}\) and \(f g\) is differentiable. Then, we have
\(\Gamma(\gamma+1) \int_{a}^{b} s^{\alpha-1} f(s)_{i} T_{M}^{\alpha, \gamma} g(s) d s=\left.f(t) g(t)\right|_{a} ^{b}-\Gamma(\gamma+1) \int_{a}^{b} s^{\alpha-1} g(s)_{i} T_{M}^{\alpha, \gamma} f(s) d s\).
The \(L_{\alpha, \gamma}^{2}(0, \pi)\) is a Hilbert space with inner product
\[
(y, z)=\int_{0}^{\pi} y^{T}\left(x, \lambda_{1}\right) z\left(x, \lambda_{2}\right) d_{\alpha, \gamma} x
\]
where \(y^{T}=\left(y_{1}, y_{2}\right)\) and \(d_{\alpha, \gamma} x=\Gamma(\gamma+1) x^{\alpha-1} d x\).
In the next section, we will analyze the Dirac systems in terms of the MTD and we are able to obtain general representations of solutions that involve parameters \(\alpha\) and \(\gamma\). Additionally, using the MTD approach, we can also present asymptotic formulas for eigen-vector-functions and eigenvalues. The general results which found in main results correspond to classical Dirac systems when \(\alpha=1\) and \(\gamma=1\).

\section*{3. Main Results}

Let us consider Dirac system containing M-tuncated derivative as follows:
\[
\begin{align*}
& { }_{i} T_{M}^{\alpha, \gamma} y_{2}(x)+p(x) y_{1}(x)=\lambda y_{1}(x), 0<\alpha \leq 1, x \in[0, \pi]  \tag{3}\\
- & { }_{i} T_{M}^{\alpha, \gamma} y_{1}(x)+r(x) y_{2}(x)=\lambda y_{2}(x),
\end{align*}
\]
where \({ }_{i} T_{M}^{\alpha, \gamma}\) is MTD operator, \(p(x)\) and \(r(x)\) are continuous and real-valued functions on \([0, \pi], y(x)\) is \(2 \alpha\)-continuously differentiable on \([0, \pi],{ }_{i} T_{M}^{\alpha, \gamma} y(x)\) is continuous on \([0, \pi]\). Deal with the system (3) subject to boundary conditions
\[
\begin{align*}
& y_{1}(0) \sin a+y_{2}(0) \cos a=0,  \tag{4}\\
& y_{1}(\pi) \sin b+y_{2}(\pi) \cos b=0, \tag{5}
\end{align*}
\]
where \(a\) and \(b\) are real constants.
Let symbolize the solution of 3 by \(\varphi(x, \lambda)=\binom{\varphi_{1}(x, \lambda)}{\varphi_{2}(x, \lambda)}\) satisfying the following initial conditions
\[
\begin{equation*}
\varphi_{1}(0, \lambda)=\cos a, \varphi_{2}(0, \lambda)=-\sin a \tag{6}
\end{equation*}
\]

Theorem 2. Let \(\lambda_{1}\) and \(\lambda_{2}\) be two distinct eigenvalues of the problem (3) - (5). Then the corresponding eigen-vector-functions \(y\left(x, \lambda_{1}\right)\) and \(z\left(x, \lambda_{2}\right)\) are orthogonal on \(L_{\alpha, \gamma}^{2}(0, \pi)\) Hilbert space, that is,
\[
\begin{equation*}
\int_{0}^{\pi} y^{T}\left(x, \lambda_{1}\right) z\left(x, \lambda_{2}\right) d_{\alpha, \gamma} x=0, \quad \lambda_{1} \neq \lambda_{2} \tag{7}
\end{equation*}
\]

Proof. Since the \(y\left(x, \lambda_{1}\right)\) and \(z\left(x, \lambda_{2}\right)\) satisfy the system (3), we have
\[
\begin{aligned}
{ }_{i} T_{M}^{\alpha, \gamma} y_{2}\left(x, \lambda_{1}\right)+p(x) y_{1}\left(x, \lambda_{1}\right) & =\lambda_{1} y_{1}\left(x, \lambda_{1}\right), \\
-{ }_{i} T_{M}^{\alpha, \gamma} y_{1}\left(x, \lambda_{1}\right)+r(x) y_{2}\left(x, \lambda_{1}\right) & =\lambda_{1} y_{2}\left(x, \lambda_{1}\right), \\
{ }_{i} T_{M}^{\alpha, \gamma} z_{2}\left(x, \lambda_{2}\right)+p(x) z_{1}\left(x, \lambda_{2}\right) & =\lambda_{2} z_{1}\left(x, \lambda_{2}\right) \\
-{ }_{i} T_{M}^{\alpha, \gamma} z_{1}\left(x, \lambda_{2}\right)+r(x) z_{2}\left(x, \lambda_{2}\right) & =\lambda_{2} z_{2}\left(x, \lambda_{2}\right) .
\end{aligned}
\]

If we multiply these equations by \(z_{1}\left(x, \lambda_{2}\right), z_{2}\left(x, \lambda_{2}\right),-y_{1}\left(x, \lambda_{1}\right)\) and \(-y_{2}\left(x, \lambda_{1}\right)\), respectively, and sum together, we get
\[
\begin{aligned}
& \left(\lambda_{1}-\lambda_{2}\right)\left(z_{1}\left(x, \lambda_{2}\right) y_{1}\left(x, \lambda_{1}\right)+z_{2}\left(x, \lambda_{2}\right) y_{2}\left(x, \lambda_{1}\right)\right) \\
& \quad={ }_{i} T_{M}^{\alpha, \gamma}\left\{z_{1}\left(x, \lambda_{2}\right) y_{2}\left(x, \lambda_{1}\right)-z_{2}\left(x, \lambda_{2}\right) y_{1}\left(x, \lambda_{1}\right)\right\}
\end{aligned}
\]

Applying the integral \({ }_{M} I_{0}^{\alpha, \gamma}\) from 0 to \(\pi\) on both side of the last equality, one can find
\(\left(\lambda_{1}-\lambda_{2}\right) \int_{0}^{\pi} y^{T}\left(x, \lambda_{1}\right) z\left(x, \lambda_{2}\right) d_{\alpha, \gamma} x=\left.\left(z_{1}\left(x, \lambda_{2}\right) y_{2}\left(x, \lambda_{1}\right)-z_{2}\left(x, \lambda_{2}\right) y_{1}\left(x, \lambda_{1}\right)\right)\right|_{0} ^{\pi}\).

By virtue of boundary conditions (4) and (5), one can obtain
\[
\left(\lambda_{1}-\lambda_{2}\right) \int_{0}^{\pi} y_{\lambda_{1}}^{T}(x) z_{\lambda_{2}}(x) d_{\alpha, \gamma} x=0
\]

Theorem 3. All eigenvalues of the problem defined by (3) - (5) are real.
Proof. Let \(\lambda_{1}=a+i b\) be an eigenvalue with eigenfunction \(y\left(x, \lambda_{1}\right)\). Since \(p(x)\) and \(r(x)\) real-valued functions, \(\lambda_{2}=\bar{\lambda}_{1}=a-i b\) is also an eigenvalue with the eigenfunctions \(\bar{y}\left(x, \lambda_{2}\right)\). By considering Theorem 2, we have
\[
\begin{aligned}
(\lambda-\bar{\lambda}) \int_{0}^{\pi} y^{T}\left(x, \lambda_{1}\right) \bar{y}\left(x, \lambda_{2}\right) d_{\alpha, \gamma} x & =0 \\
(\lambda-\bar{\lambda}) \int_{0}^{\pi}\left\{y_{1}^{2}\left(x, \lambda_{1}\right)+y_{2}^{2}\left(x, \lambda_{1}\right)\right\} d_{\alpha, \gamma} x & =0
\end{aligned}
\]
and since \(y(x) \neq 0\), we have \(\lambda=\bar{\lambda}\).
Theorem 4. The solution of the system (3) satisfying the initial conditions (6) provides the following integral equation system,
\[
\begin{align*}
\varphi_{1}(x, \lambda)= & \cos \left(\frac{\lambda \Gamma(\gamma+1) x^{\alpha}}{\alpha}-a\right)-\int_{0}^{x} \sin \left(\lambda \Gamma(\gamma+1)\left(\frac{t^{\alpha}-x^{\alpha}}{\alpha}\right)\right) p(t) \varphi_{1}(t, \lambda) d_{\alpha, \gamma} t \\
& +\int_{0}^{x} \cos \left(\lambda \Gamma(\gamma+1)\left(\frac{t^{\alpha}-x^{\alpha}}{\alpha}\right)\right) r(t) \varphi_{2}(t, \lambda) d_{\alpha, \gamma} t  \tag{8}\\
\varphi_{2}(x, \lambda)= & \sin \left(\frac{\lambda \Gamma(\gamma+1) x^{\alpha}}{\alpha}-a\right)-\int_{0}^{x} \cos \left(\lambda \Gamma(\gamma+1)\left(\frac{t^{\alpha}-x^{\alpha}}{\alpha}\right)\right) p(t) \varphi_{1}(t, \lambda) d_{\alpha, \gamma} t \\
& -\int_{0}^{x} \sin \left(\lambda \Gamma(\gamma+1)\left(\frac{t^{\alpha}-x^{\alpha}}{\alpha}\right)\right) r(t) \varphi_{2}(t, \lambda) d_{\alpha, \gamma} t \tag{9}
\end{align*}
\]

Proof. By using the variation of parameters method given in 17, we express the representation of the solutions as follow:
\[
\begin{aligned}
\varphi_{1}(x, \lambda) & =-c_{1}(x) y_{1}(x)+c_{2}(x) y_{2}(x) \\
\varphi_{2}(x, \lambda) & =c_{1}(x) y_{2}(x)+c_{2}(x) y_{1}(x)
\end{aligned}
\]
where
\[
\begin{aligned}
& c_{1}(x)=-\int_{0}^{x}\left(p(t) y_{1}(t) \varphi_{1}(t, \lambda)-r(t) y_{2}(t) \varphi_{2}(t, \lambda)\right) d_{\alpha, \gamma} t+c_{1}, \\
& c_{2}(x)=\int_{0}^{x}\left(p(t) y_{1}(t) \varphi_{2}(t, \lambda)+r(t) y_{2}(t) \varphi_{1}(t, \lambda)\right) d_{\alpha, \gamma} t+c_{2}, \\
& y_{1}(x)=\sin \frac{\lambda \Gamma(\gamma+1) x^{\alpha}}{\alpha} \text { and } y_{2}(x)=\cos \frac{\lambda \Gamma(\gamma+1) x^{\alpha}}{\alpha} .
\end{aligned}
\]

If we benefit from the initial conditions (6), it can be easily seen (8) and (9).
Theorem 5. As \(|\lambda| \rightarrow \infty\), the estimates are provided as follows:
\[
\left.\begin{array}{c}
\varphi_{1}(x, \lambda)=\cos (\xi(x, \lambda)-a)+O\left(\frac{1}{\lambda}\right) \\
\varphi_{2}(x, \lambda)=\sin (\xi(x, \lambda)-a)+O\left(\frac{1}{\lambda}\right) \\
\frac{\partial \varphi_{1}(x, \lambda)}{\partial \lambda}=-\Gamma(\gamma+1) \frac{x^{\alpha}}{\alpha} \sin (\xi(x, \lambda)-a)+O(1) \\
\frac{\partial \varphi_{2}(x, \lambda)}{\partial \lambda}= \tag{13}
\end{array}\right) \Gamma(\gamma+1) \frac{x^{\alpha}}{\alpha} \cos (\xi(x, \lambda)-a)+O(1), ~ \$
\]
for \(0 \leq x \leq \pi\) where
\[
\begin{equation*}
\xi(x, \lambda)=\frac{\lambda \Gamma(\gamma+1)}{\alpha} x^{\alpha}+\frac{1}{2} \int_{0}^{x}(p(t)+r(t)) d_{\alpha, \gamma} t \tag{14}
\end{equation*}
\]

Proof. Let us introduce by \(\varphi(x, \lambda)\) the solution of the system (3) satisfying the initial conditions (6). If the problem (3), (6) is considered for \(p(x)=r(x) \equiv 0\), the solution of this problem stand for \(\psi(x, \lambda)\). Therby, one can easily obtain that
\[
\begin{gather*}
\psi_{1}(x, \lambda)=\cos \left(\frac{\lambda \Gamma(\gamma+1)}{\alpha} x^{\alpha}-a\right)  \tag{15}\\
\psi_{2}(x, \lambda)=\sin \left(\frac{\lambda \Gamma(\gamma+1)}{\alpha} x^{\alpha}-a\right) \tag{16}
\end{gather*}
\]

If the solution of the problem (3), (6) is applied to the transformation matrix operator, we have 22
\[
\begin{equation*}
\varphi(x, \lambda)=R(x) \psi(x, \lambda)+\int_{0}^{x} K(x, s) \psi(s, \lambda) d_{a, \gamma} s \tag{17}
\end{equation*}
\]
in here \(R(x)\) and \(K(x, s)\) are matrices of second-order that can be continuously differentiated twice,
\[
R(x)=\left(\begin{array}{cc}
\gamma(x) & \beta(x)  \tag{18}\\
-\beta(x) & \gamma(x)
\end{array}\right)
\]
and \(\gamma(x)\) and \(\beta(x)\) can be computed as below:
\[
\begin{aligned}
& \gamma(x)=\cos \left(\frac{1}{2} \int_{0}^{x}(p(t)+r(t)) d_{\alpha, \gamma} t\right) \\
& \beta(x)=-\sin \left(\frac{1}{2} \int_{0}^{x}(p(t)+r(t)) d_{\alpha, \gamma} t\right)
\end{aligned}
\]
for \(\kappa=1\). Thereby considering by (17) and (18), we find the formulas
\[
\left.\left.\begin{array}{rl}
\varphi_{1}(x, \lambda)= & \cos (
\end{array}(x, \lambda)-a\right)+\int_{0}^{x} K_{11}(x, s) \cos \left(\frac{\lambda \Gamma(\gamma+1)}{\alpha} s^{\alpha}-a\right) d_{\alpha, \gamma} s\right)
\]
where \(K_{i j}(x, s)\) are the components of the matrix \(K(x, s)\) for \(i, j=1,2\) from (17). To gain the asymptotics in (10) and (11), it is enough to integrate by parts the integrals including in (19), because of the differentiability of the functions \(K_{i j}(x, s)\). In a similar manner, if we differentiate (19) in terms of \(\lambda\), we obtain the asymptotics in (12) and (13).

Additionally, we demonstrate the asymptotic behaviors of the eigenvalues using the MTD approach, enabling us to observe how the formulas change as \(\alpha\) and \(\gamma\) vary.

Theorem 6. The eigenvalues of the problem outlined by equations (3) to (5) in their asymptotic forms are given as follows:
\[
\begin{equation*}
\lambda_{ \pm n}=\frac{\alpha}{\Gamma(\gamma+1) \pi^{\alpha}}( \pm n \pi+c)+O\left(\frac{1}{n}\right) \tag{20}
\end{equation*}
\]
where
\[
c=a-b-\frac{1}{2} \int_{0}^{x}(p(t)+r(t)) d_{\alpha, \gamma} t
\]

Proof. The eigenvalues of the given problem overlap with the roots of the characteristic function
\[
\triangle(\lambda)=\varphi_{1}(\pi, \lambda) \sin b+\varphi_{2}(\pi, \lambda) \cos b
\]

If we put asymptotics of the eigen-vector-functions \(\varphi_{1}(\pi, \lambda)\) and \(\varphi_{2}(\pi, \lambda)\) from the estimates (11) into \(\triangle(\lambda)\), we obtain
\[
\cos (\xi(x, \lambda)-a) \sin b+\sin (\xi(x, \lambda)-a) \cos b+O\left(\lambda^{-1}\right)=0
\]

After some calculation with the aid of trigonometric functions, we reach
\[
\begin{equation*}
\sin \left(\frac{\lambda \Gamma(\gamma+1) \pi^{\alpha}}{\alpha}+c\right)+O\left(\lambda^{-1}\right)=0 \tag{21}
\end{equation*}
\]

It is clearly seen that the equation (21), for large \(|\lambda|\), has solutions in the form
\[
\frac{\lambda \Gamma(\gamma+1) \pi^{\alpha}}{\alpha}+c=n \pi+\delta_{n}
\]
it is obvious that \(\sin \delta_{n}=O\left(n^{-1}\right)\), i.e. \(\delta_{n}=O\left(n^{-1}\right)\). Therefore the asymptotic formula for eigenvalues is obtained in (20).

Theorem 7. The asymptotic formula for the norming constants is given by
\[
\rho_{n}=\sqrt{\frac{\pi^{\alpha} \Gamma(\gamma+1)}{\alpha}}+O\left(\frac{1}{n}\right)
\]

Proof. By utilizing the asymptotic formula for eigenvalues given in (20), we can reobtain the asymptotics for eigen-vector-functions as follows:
\[
\begin{align*}
& \varphi_{1}\left(x, \lambda_{n}\right)=\cos \left(\xi_{n}-a\right)+O\left(n^{-1}\right)  \tag{22}\\
& \varphi_{2}\left(x, \lambda_{n}\right)=\sin \left(\xi_{n}-a\right)+O\left(n^{-1}\right) \tag{23}
\end{align*}
\]
where \(\xi\left(x, \lambda_{n}\right)=\xi_{n}=\frac{\lambda_{n} \Gamma(\gamma+1)}{\alpha} x^{\alpha}+\frac{1}{2} \int_{0}^{x}(p(t)+r(t)) d_{\alpha, \gamma} t\).
To reach at the asymptotic expression for the norming constants, take in consideration the following integral
\[
\begin{aligned}
\rho_{n}^{2} & =\int_{0}^{\pi}\left\{\varphi_{1}^{2}\left(x, \lambda_{n}\right)+\varphi_{2}^{2}\left(x, \lambda_{n}\right)\right\} d_{\alpha, \gamma} x \\
& =\int_{0}^{\pi}\left\{\cos ^{2}\left(\xi_{n}-a\right) x+\sin ^{2}\left(\xi_{n}-a\right)\right\} d_{\alpha, \gamma} x+O\left(\frac{1}{n}\right) \\
& =\frac{\pi^{\alpha} \Gamma(\gamma+1)}{\alpha}+O\left(\frac{1}{n}\right)
\end{aligned}
\]

Hence, the proof is completed.

Theorem 8. Asymptotic expression of the normalized vector-valued eigenfunctions is given in the form,
\[
\widetilde{\varphi}\left(x, \lambda_{n}\right)=\binom{\sqrt{\frac{\alpha}{\pi^{\alpha} \Gamma(\gamma+1)}} \cos \left(\xi_{n}-a\right)+O\left(n^{-1}\right)}{\sqrt{\frac{\alpha}{\pi^{\alpha} \Gamma(\gamma+1)}} \sin \left(\xi_{n}-a\right)+O\left(n^{-1}\right)}
\]

Proof. The proof can be easily seen with the help of Theorem 7.

\section*{4. Illustrative Results}

In the current section, the representation of the solutions \(y_{1}(x)\) and \(y_{2}(x)\) for Dirac equation is offered by means of MTD under different orders of \(\alpha\), different potentials and different values of \(\lambda\). If the values of \(\alpha\) increases while the value of \(\gamma\) is constant, Figure 1 (a) and (b) have showed a right-sided shift for the solution curves. If the values of \(\gamma\) increases while the value of \(\alpha\) is constant, Figure 2 (a) and (b) have showed a smaller right-sided shift in the solutions than Figure 1. Figure 3 demonstrates the acting for the solutions when \(q=0,1,2,3\). Also, the roots of the characteristic function are computed detailed under different values of \(\alpha\) in Table 1. If one pays attention to Figure 4 (a), (b) and (c), it can be easily seen that the value of \(\alpha\) increases which is equal to \(0.1,0.3,0.5\), respectively, the frequency of the oscillation interval decreases. That is as the value of \(\alpha\) decreases the number of eigenvalues of considered problem increases. Thereby \(\alpha\) is changed the mobility of the solutions curves increases and it provides important advantage in applications of spectral analysis. Lastly, in Figure 5 (a) and (b), the graphs of eigenfunctions corresponding to different eigenvalues were plotted according to the changing the value of \(\alpha\) and \(\gamma\), respectively. Also Figure 5 (b) shows that eigenfunctions overlap for different values of \(\gamma\). The main purpose in drawing graphs with different values is that one can observe the behavior of representations of solutions curves for Dirac equation in light of MTD. Also the approximate eigenvalues are given for different orders of \(\alpha\) and \(\gamma\) in Table 1. Assume that \(a=1, b=\frac{\pi}{4}\) for all figures.

Table 1. The roots of \(\triangle(\lambda)\) for \(x=\pi\)
\begin{tabular}{|c|c|c|c|c|c|}
\hline\(\alpha\) & \(\lambda_{1}\) & \(\lambda_{2}\) & \(\gamma\) & \(\lambda_{1}\) & \(\lambda_{2}\) \\
\hline 0.1 & -0.2945 & 0.0215 & 0.1 & -0.8679 & 0.0636 \\
\hline 0.3 & -0.7028 & 0.0515 & 0.3 & -0.9200 & 0.0674 \\
\hline 0.5 & -0.9316 & 0.0683 & 0.5 & -0.9316 & 0.0683 \\
\hline 0.7 & -1.0374 & 0.0760 & 0.7 & -0.9087 & 0.0666 \\
\hline 0.9 & -1.0609 & 0.0777 & 0.9 & -0.8585 & 0.0629 \\
\hline 0.99 & -1.0527 & 0.0771 & 0.99 & -0.8291 & 0.0607 \\
\hline
\end{tabular}


Figure 1. Comparative analysis for different orders of \(\alpha, \lambda=\) \(10, p(x)=r(x)=0, \gamma=0.5\)
\(-\gamma=0.1-\gamma=0.3-\gamma=0.5-\gamma=0.7-\gamma=0.9-\gamma=0.1-\gamma=0.3-\gamma=0.5-\gamma=0.7-\gamma=0.9\)


Figure 2. Comparative analysis for different orders of \(\gamma, \lambda=\) \(10, p(x)=r(x)=0, \alpha=0.5\).


Figure 3. Comparative analysis for different values of the potentials, \(\lambda=10, \quad p(x)=r(x)=q, \gamma=0.5, \alpha=0.5\)



Figure 4. Comparisons of the roots of the characteristic function under different orders, \(\lambda=10, p(x)=r(x)=0, \gamma=0.5\)


Figure 5. Comparisons of the eigenfunctions benefit from Table 1 1 under different order of \(\alpha\) and \(\gamma\)

\section*{5. Conclusion}

In here, we analyzed spectral structure of Dirac systems which has been studied by Levitan and Sargsjan 22 for integer order case in light of MTD. For onedimensional Dirac operator in sense of MTD, its fundamental spectral theory is given systematically and behaviours of eigen-vector-functions are observed with graphics under different orders, potentials and eigenvalues. We obtain the representations of the solutions and asymptotics for the norming constants, the eigenvalues, eigen-vector-functions, and the normalized eigen-vector-functions. To gain these important results, certain calculations like variation of parameters method, Leibniz rule, and so forth are made in sense of MTD. The most important advantage of MTD is that this definition offers the features of the integer-order calculus. MTD give us the change to examine derivatives of infinite order . Also, we give comparative analysis of the solutions by graphs with different orders \(\alpha\) and \(\gamma\), different eigenvalues and different potentials. Thereby, we observe the behaviours of the mobility of the solutions. Thus, we have supplied a large amount of spectral theory for the considered problem in terms of MTD.

Author Contribution Statements As the sole author, the work is entirely the author's own.

Declaration of Competing Interests The author declares that there is no competing interest regarding the publication of this paper.

Acknowledgements The author would like to thank the editor and the reviewers for their valuable comments and suggestions which resulted in substantial improvement in the presentation of the paper.

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Volume 73, Number 1, Pages 274-284 (2024)
DOI:10.31801/cfsuasmas. 1313970
ISSN 1303-5991 E-ISSN 2618-6470
Research Article; Received: June 13, 2023; Accepted: October 31, 2023

\title{
APPLICATION OF THE GKM TO SOME NONLINEAR PARTIAL EQUATIONS
}

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\begin{abstract}
In this manuscript, the strain wave equation, which plays an important role in describing different types of wave propagation in microstructured solids and the \((2+1)\) dimensional Bogoyavlensky Konopelchenko equation, is defined in fluid mechanics as the interaction of a Riemann wave propagating along the \(y\)-axis and a long wave propagating along the \(x\)-axis, were studied. The generalized Kudryashov method (GKM), which is one of the solution methods of partial differential equations, was applied to these equations for the first time. Thus, a series of solutions of these equations were obtained. These found solutions were compared with other solutions. It was seen that these solutions were not shown before and were presented for the first time in this study. The new solutions of these equations might have been useful in understanding the phenomena in which waves are governed by these equations. In addition, 2D and 3D graphs of these solutions were constructed by assigning certain values and ranges to them.
\end{abstract}

\section*{1. Introduction}

Nonlinear evolution equations (NLEEs) have been utilized to make mathematical models of encountered problems in various scientific circles. A number of solution methods have been developed by various scientists to solve NLEEs, which have very important areas of use \(1-10\). In this study, one of these methods, GKM, has been taken into consideration and applied to the strain wave and ( \(2+1\) )-dimensional Bogoyavlensky-Konopelchenko (BK) equations.

\footnotetext{
2020 Mathematics Subject Classification. 35A25,35C07,35C08.
Keywords. Generalized Kudryashov method, strain wave equation, (2+1)-dimensional Bogoyavlensky-Konopelchenko equation.

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Strain wave equation is given as 11:
\[
\begin{align*}
u_{t t}- & u_{x x}-\epsilon \alpha_{1}\left(u^{2}\right)_{x x}-k \alpha_{2} u_{x x t}+\delta \alpha_{3} u_{x x x x} \\
& \quad-\left(\delta \alpha_{4}+k^{2} \alpha_{7}\right) u_{x x t t}+k \delta\left(\alpha_{5} u_{x x x x t}+\alpha_{6} u_{x x t t t}\right)=0 \tag{1}
\end{align*}
\]
where \(u(x, t)\) is the micro-strain wave function. \(\epsilon\) indicates elastic strain, \(\delta\) shows the elastic stresses and the rate between the wavelength and size of the microstructure, \(k\) reflects the dissipative effect and \(\alpha_{1}, \alpha_{2}, \alpha_{3}, \alpha_{4}, \alpha_{5}, \alpha_{6}, \alpha_{7}\) are arbitrary constants. Assuming \(\delta=O(\epsilon)\) on Eq. (11), an equilibration takes place between dispersion and nonlinearity. If \(k=0\) is selected in this equation, the undistributed state of the micro-stress wave is obtained. In this way, the following equation for the bi-dispersion in microstructured solids is obtained 12-16:
\[
\begin{equation*}
u_{t t}-u_{x x}-\epsilon\left(\alpha_{1}\left(u^{2}\right)_{x x}-\alpha_{3} u_{x x x x}+\alpha_{4} u_{x x t t}\right)=0 \tag{2}
\end{equation*}
\]

Recently, the solutions of strain wave equation investigated by various researchers with different methods. Seadawy et al. used the modified extended mapping method for strain wave equation 11. Ayati et al. applied the functional variable method and Kudryashov method to strain wave equation 12 . Arshad et al. practiced the modified direct algebraic method to strain wave equation 13 . Gao et al. used the F-expansion method for strain wave equation 14 . Irshad et al. practiced the generalized Jacobi elliptic function method to strain wave equation 15. Kumar et al. used the generalized exponential rational function method for strain wave equation 16. Joseph implemented the new rational F-expansion method to strain wave equation 17 .
\((2+1)\)-dimensional BK equation is given as 18:
\[
\begin{equation*}
u_{x t}+h_{1} u_{x x x x}+h_{2} u_{x x x y}+h_{3} u_{x x} u_{x}+h_{4}\left(u_{x y} u_{x}+u_{x x} u_{y}\right)=0 \tag{3}
\end{equation*}
\]
where \(h_{1}, h_{2}, h_{3}\) and \(h_{4}\) are arbitrary constants. If \(h_{1}=a, h_{2}=\beta, h_{3}=6 a, h_{4}=4 \beta\) values are selected for the \(h_{1}, h_{2}, h_{3}, h_{4}\) constants in Eq. (3), Eq. (3) can be written as.
\[
\begin{equation*}
u_{x t}+\alpha u_{x x x x}+\beta u_{x x x y}+6 \alpha u_{x x} u_{x}+4 \beta u_{x y} u_{x}+4 \beta u_{x x} u_{y}=0 \tag{4}
\end{equation*}
\]

The resulting Eq. (4) is handled as a two-dimensional generalization of the KdV equation, and under favorable conditions, it can be converted to the KdV equation 19. This equation provides the Calogero-Bogoyavlensky-Schiff equation for \(\alpha=0\) and is also defined as the interplay of a Riemann wave spreading along the y -axis and a long wave spreading along the \(x\)-axis in fluid mechanics 20, 21. For Eq. (4) \(u_{y}=v_{x}\) is transformed and integrated, and the following equation is found:
\[
\begin{equation*}
u_{t}+\alpha u_{x x x}+\beta v_{x x x}+3 \alpha\left(u_{x}\right)^{2}+4 \beta u_{x} v_{x}=0 \tag{5}
\end{equation*}
\]

Accordingly, Eq. (4) can be expressed as a system as follows:
\[
\begin{array}{r}
u_{t}+\alpha u_{x x x}+\beta v_{x x x}+3 \alpha\left(u_{x}\right)^{2}+4 \beta u_{x} v_{x}=0 \\
u_{y}=v_{x} \tag{6}
\end{array}
\]

When looking at the past works on (2+1)-dimensional BK equation. Zhou et al. gave based on its bilinear form, the N th-order breather solutions of the \((2+1)\) dimensional generalized BK equation 21. Ray got infinitesimal generators of (2+1)-dimensional BK equation by using Lie group analysis method and investigated symmetry analysis and similarity reduction of \((2+1)\)-dimensional BK equation 1822 . Chen and Ma obtained the symbolic solutions of the ( \(2+1\) )-dimensional BK equation that own a Hirota bilinear form 23.

The purpose of this article is to detect soliton solutions of strain wave equation and (2+1)-dimensional BK equation using GKM 24 27. First of all, the features of GKM, which is the method we used in our study, are explained. Subsequently, some soliton solutions of the strain wave equation and \((2+1)\)-dimensional BK equation were found using this method.

\section*{2. Analysis of the Method}

Consider a general nonlinear partial differential equation for a function \(v\) that depends on three variables, as follows:
\[
\begin{equation*}
K\left(v, v_{t}, v_{y}, v_{x}, v_{x x}, \ldots\right)=0 \tag{7}
\end{equation*}
\]

Step 1: First, the traveling wave transform is discussed in the following form;
\[
\begin{equation*}
v(x, y, t)=v(\eta), \eta=x+y-m t \tag{8}
\end{equation*}
\]

Eq. (7) is transformed into an ordinary differential equation using the transformations in Eq. (8) as follows:
\[
\begin{equation*}
L\left(t, y, x, v, v^{\prime}, v^{\prime \prime}, \cdots\right)=0 \tag{9}
\end{equation*}
\]
where superscripts demonstrate ordinary derivatives according \(\eta\)
Step 2: Assume that the solutions of Eq. (9) are treated as follows:
\[
\begin{equation*}
v(\eta)=\frac{\sum_{i=0}^{\sigma} a_{i} Q^{i}(\eta)}{\sum_{j=0}^{\rho} b_{j} Q^{j}(\eta)}=\frac{P[Q(\eta)]}{S[Q(\eta)]} \tag{10}
\end{equation*}
\]
where \(Q\) is \(\frac{1}{1 \pm e^{\eta}}\). It is stated that \(Q\) is the solution of the following equation
\[
\begin{equation*}
Q_{\eta}=Q^{2}-Q \tag{11}
\end{equation*}
\]

Step 3: The solution of Eq. (9) is sought according to this method as follows:
\[
\begin{equation*}
v(\eta)=\frac{a_{0}+a_{1} Q+a_{2} Q^{2}+\cdots+a_{\sigma} Q^{\sigma}}{b_{0}+b_{1} Q+b_{2} Q^{2}+\cdots+b_{\rho} Q^{\rho}} \tag{12}
\end{equation*}
\]

The values of \(\sigma\) and \(\rho\) in Eq. 10 can be determined through the homogeneous balance principle. For this, a balance is established between the highest-order
derivative and the highest-order nonlinear term in Eq. (9).
Step 4: Eq. 10 is inserted into Eq. (9). Thus, a polynomial \(R(Q)\) of \(Q\) is obtained. Thereafter all coefficients of \(R(Q)\) are set equal to zero, to obtain a system of algebraic equations. Solving the resulting system determines \(c\) and the coefficients \(a_{0}, a_{1}, a_{2}, \ldots, a_{\sigma}, b_{0}, b_{1}, b_{2}, \ldots, b_{\rho}\). Finally, the soliton solutions of Eq. (9) are obtained.

\section*{3. Application of GKM to the equations}

Example 1. Initially, the following transformation is considered.
\[
\begin{equation*}
u(x, t)=u(\eta), \eta=x-c t \tag{13}
\end{equation*}
\]

Substituting Eq. (13) into Eq. (2) yields the following equation.
\[
\begin{equation*}
\left(c^{2}-1\right) u-\epsilon \alpha_{1} u^{2}+\epsilon\left(\alpha_{3}-c^{2} \alpha_{4}\right) u^{\prime \prime}=0 \tag{14}
\end{equation*}
\]

If the balance principle is applied to Eq. (14), the following equation is obtained
\[
\sigma=\rho+2
\]

If \(\rho=1\), then \(\sigma=3\). Thus the following equations are found.
\[
\begin{equation*}
u(\eta)=\frac{a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}}{b_{0}+b_{1} Q} \tag{15}
\end{equation*}
\]
\[
\begin{aligned}
u^{\prime}(\eta)= & \left(Q^{2}-Q\right) \\
\times & {\left[\frac{\left(a_{1}+2 a_{2} Q+3 a_{3} Q^{2}\right)\left(b_{0}+b_{1} Q\right)-b_{1}\left(a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}\right)}{\left(b_{0}+b_{1} Q\right)^{2}}\right] } \\
u^{\prime \prime}(\eta)= & \frac{\left(Q^{2}-Q\right)(2 Q-1)}{\left(b_{0}+b_{1} Q\right)^{2}} \\
& \times\left[\left(a_{1}+2 a_{2} Q+3 a_{3} Q^{2}\right)\left(b_{0}+b_{1} Q\right)-b_{1}\left(a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}\right)\right] \\
& +\frac{\left(Q^{2}-Q\right)^{2}}{\left(b_{0}+b_{1} Q\right)^{3}}\left[\left(2 a_{2}+6 a_{3} Q\right)\left(b_{0}+b_{1} Q\right)^{2}-2 b_{1}\left(a_{1}+2 a_{2} Q+3 a_{3} Q^{2}\right)\left(b_{0}+b_{1} Q\right)\right] \\
& +\frac{\left(Q^{2}-Q\right)^{2}}{\left(b_{0}+b_{1} Q\right)^{3}}\left[2 b_{1}^{2}\left(a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}\right)\right]
\end{aligned}
\]

The soliton solutions of the strain wave equation are obtained in different cases as follows;

Case 1.
\[
\begin{gathered}
a_{0}=0, a_{1}=\frac{6 b_{0}\left(\alpha_{3}-\alpha_{4}\right)}{\alpha_{1}\left(-1+\epsilon \alpha_{4}\right)}, a_{3}=\frac{6 b_{1}\left(-\alpha_{3}+\alpha_{4}\right)}{\alpha_{1}\left(-1+\epsilon \alpha_{4}\right)} \\
a_{2}=\frac{6\left(-b_{0}+b_{1}\right)\left(\alpha_{3}-\alpha_{4}\right)}{\alpha_{1}\left(-1+\epsilon \alpha_{4}\right)}, c=-\frac{\sqrt{-1+\epsilon \alpha_{3}}}{\sqrt{-1+\epsilon \alpha_{4}}}
\end{gathered}
\]

By substituting the above equalities into Eq. (15), the following solution of Eq. (2) is found.
\[
\begin{equation*}
u_{1}(x, t)=\frac{3\left(\alpha_{3}-\alpha_{4}\right)}{\left(1+\cosh \left[x+t \frac{\sqrt{-1+\epsilon \alpha_{3}}}{\sqrt{-1+\epsilon \alpha_{4}}}\right]\right) \alpha_{1}\left(-1+\epsilon \alpha_{4}\right)} . \tag{16}
\end{equation*}
\]


Figure 1. 3D and 2D plots of \(u_{1}(x, t)\) solution.
Case 2.
\[
\begin{array}{r}
a_{0}=\frac{b_{0}\left(\alpha_{3}-\alpha_{4}\right)}{\alpha_{1}\left(1+\epsilon \alpha_{4}\right)}, a_{1}=\frac{\left(-6 b_{0}+b_{1}\right)\left(\alpha_{3}-\alpha_{4}\right)}{\alpha_{1}\left(1+\epsilon \alpha_{4}\right)}, \\
a_{2}=\frac{6\left(b_{0}-b_{1}\right)\left(\alpha_{3}-\alpha_{4}\right)}{\alpha_{1}\left(1+\epsilon \alpha_{4}\right)}, \\
a_{3}=\frac{6 b_{1}\left(\alpha_{3}-\alpha_{4}\right)}{\alpha_{1}\left(1+\epsilon \alpha_{4}\right)}, c=\frac{\sqrt{1+\epsilon \alpha_{3}}}{\sqrt{1+\epsilon \alpha_{4}}}
\end{array}
\]

By substituting the above equalities into Eq. (15), the following solution of Eq. (2) is found.
\[
\begin{equation*}
u_{2}(x, t)=\frac{\left(-2+\cosh \left[x-t \frac{\sqrt{1+\epsilon \alpha_{3}}}{\sqrt{1+\epsilon \alpha_{4}}}\right]\right)\left(\alpha_{3}-\alpha_{4}\right)}{\left(1+\cosh \left[x-t \frac{\sqrt{1+\epsilon \alpha_{3}}}{\sqrt{1+\epsilon \alpha_{4}}}\right]\right) \alpha_{1}\left(1+\epsilon \alpha_{4}\right)} . \tag{17}
\end{equation*}
\]


Figure 2. 3 D and 2 D plots of \(u_{2}(x, t)\) solution.

Example 2. First, he following transformation is taken into account.
\[
\begin{equation*}
u(x, y, t)=u(\eta), v(x, y, t)=v(\eta), \eta=k x+m y-c t \tag{18}
\end{equation*}
\]

Substituting Eq. (18) into system (6) yields the following equation.
\[
\begin{equation*}
-c u^{\prime}+\left(\alpha k^{3}+m \beta k^{2}\right) u^{\prime \prime \prime}+\left(3 \alpha k^{2}+4 m \beta k\right)\left(u^{\prime}\right)^{2}=0 . \tag{19}
\end{equation*}
\]

The following equation is obtained by transformation \(u^{\prime}=g\) in Eq. 19.
\[
\begin{equation*}
-c g+\left(\alpha k^{3}+m \beta k^{2}\right) g^{\prime \prime}+\left(3 \alpha k^{2}+4 m \beta k\right) g^{2}=0 \tag{20}
\end{equation*}
\]

As a result of applying (18) transformation to this system, \(v=\frac{m}{k} u\) equality is obtained. If the balance principle is applied to Eq. 20, the following equation is obtained.
\[
\sigma=\rho+2
\]

If \(\rho=1\), then \(\sigma=3\). Thus the following equations are found.
\[
\begin{equation*}
u(\eta)=\frac{a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}}{b_{0}+b_{1} Q} \tag{21}
\end{equation*}
\]
\[
\begin{aligned}
u^{\prime}(\eta)= & \left(Q^{2}-Q\right) \\
& \times\left[\frac{\left(a_{1}+2 a_{2} Q+3 a_{3} Q^{2}\right)\left(b_{0}+b_{1} Q\right)-b_{1}\left(a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}\right)}{\left(b_{0}+b_{1} Q\right)^{2}}\right],
\end{aligned}
\]
\[
\begin{aligned}
& u^{\prime \prime}(\eta)=\frac{\left(Q^{2}-Q\right)(2 Q-1)}{\left(b_{0}+b_{1} Q\right)^{2}} \\
& \quad \times\left[\left(a_{1}+2 a_{2} Q+3 a_{3} Q^{2}\right)\left(b_{0}+b_{1} Q\right)-b_{1}\left(a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}\right)\right] \\
& \quad+\frac{\left(Q^{2}-Q\right)^{2}}{\left(b_{0}+b_{1} Q\right)^{3}}\left[\left(2 a_{2}+6 a_{3} Q\right)\left(b_{0}+b_{1} Q\right)^{2}-2 b_{1}\left(a_{1}+2 a_{2} Q+3 a_{3} Q^{2}\right)\left(b_{0}+b_{1} Q\right)\right] \\
& \quad+\frac{\left(Q^{2}-Q\right)^{2}}{\left(b_{0}+b_{1} Q\right)^{3}}\left[2 b_{1}^{2}\left(a_{0}+a_{1} Q+a_{2} Q^{2}+a_{3} Q^{3}\right)\right]
\end{aligned} .
\]

The soliton solutions of the (2+1)-dimensional BK equation are obtained in different cases as follows;;

Case 1.
\[
\begin{aligned}
a_{0}=0, a_{1}=-\frac{a_{2}}{6}, a_{3}=-a_{2}, b_{0} & =0, c=\frac{k^{2} m \beta a_{2}}{3 a_{2}-6 k b_{1}}, \\
\alpha & =-\frac{2 m \beta\left(2 a_{2}-3 k b_{1}\right)}{3 k\left(a_{2}-2 k b_{1}\right)} .
\end{aligned}
\]

Replacing the above equations in Eq. 21, the following solution of system (6) is reached.
\[
\begin{align*}
& u_{1}(x, y, t)=\frac{a_{2}}{2 b_{1}}\left(\tanh \left[\frac{k x}{2}+\frac{m y}{2}-\frac{k^{2} m t \beta a_{2}}{6 a_{2}-12 k b_{1}}\right]-\frac{k x}{3}-\frac{m y}{3}+\frac{k^{2} m t \beta a_{2}}{9 a_{2}-18 k b_{1}}\right) .  \tag{22}\\
& v_{1}(x, y, t)=\frac{m a_{2}}{2 k b_{1}}\left(\tanh \left[\frac{k x}{2}+\frac{m y}{2}-\frac{k^{2} m t \beta a_{2}}{6 a_{2}-12 k b_{1}}\right]-\frac{k x}{3}-\frac{m y}{3}+\frac{k^{2} m t \beta a_{2}}{9 a_{2}-18 k b_{1}}\right) .
\end{align*}
\]



Figure 3. 3D and 2D plots of \(u_{1}(x, y, t)\) solution.

Case 2.
\[
\begin{gathered}
a_{0}=0, a_{1}=-\frac{k(k \alpha+m \beta) b_{1}}{3 k \alpha+4 m \beta}, a_{2}=\frac{6 k(k \alpha+m \beta) b_{1}}{3 k \alpha+4 m \beta}, \\
a_{3}=-\frac{6 k(k \alpha+m \beta) b_{1}}{3 k \alpha+4 m \beta}, b_{0}=0, c=-k^{2}(k \alpha+m \beta) .
\end{gathered}
\]

Replacing the above equations in Eq. 21], the following solution of system (6) is reached.
\[
\begin{align*}
& u_{2}(x, y, t)=-\frac{k(k \alpha+m \beta)\left(k x+m y+k^{2} t(k \alpha+m \beta)-3 \tanh \left[\frac{1}{2}\left(k x+m y+k^{2} t(k \alpha+m \beta)\right)\right]\right)}{3 k \alpha+4 m \beta} \\
& v_{2}(x, y, t)=-\frac{m(k \alpha+m \beta)\left(k x+m y+k^{2} t(k \alpha+m \beta)-3 \tanh \left[\frac{1}{2}\left(k x+m y+k^{2} t(k \alpha+m \beta)\right)\right]\right)}{3 k \alpha+4 m \beta} \tag{23}
\end{align*}
\]


Figure 4. 3D and 2D plots of \(u_{2}(x, y, t)\) solution.

\section*{4. Results and Discussion}

In this study, strain wave and (2+1)-dimensional BK equations are studied. Hyperbolic solutions for the strain wave equation and dark soliton solutions for the ( \(2+1\) )-dimensional BK equation are obtained. When these solutions are compared with previous studies in the literature, it is seen that the solutions are new and presented for the first time in this study. The graphical representations of the obtained solutions are made for the following values.
Figure 1, depicts singular kink soliton for 3D plot of solution (16) for \(\alpha_{1}=2, \alpha_{3}=3, \alpha_{4}=\) \(0.5, \epsilon=4,-25 \leq x \leq 25\) values with \(-5 \leq t \leq 5\) range and 2D plot of solution for \(t=2.5\) with these values. Figure 2, shows singular kink soliton for 3D plot of solution (17) for \(\alpha_{1}=1.5, \alpha_{3}=2, \alpha_{4}=0.2, \epsilon=1.5,-20 \leq x \leq 20\) values with \(-4 \leq t \leq 4\) range and 2D plot of solution for \(t=3\) with these values. Figure 3, represents soliton solution for 3D plot of solution (22) for \(a_{2}=2, b_{1}=1, k=0.05, m=1, \beta=1, y=1,-40 \leq x \leq 40\) values with \(-3 \leq t \leq 3\) range and 2D plot of solution for \(t=2\) with these values. Figure 4, depicts smooth soliton for 3D plot of solution (23) for \(k=1, m=0.2, \alpha=0.2, \beta=\) \(0.5, y=2,-25 \leq x \leq 25\) values with \(-5 \leq t \leq 5\) range and 2D plot of solution for \(t=3\) with these values.

\section*{5. Conclusions}

In this study, GKM was considered. GKM was applied to the strain wave equation and \((2+1)\)-dimensional BK equations. Thus, hyperbolic soliton solutions of the strain wave equation and dark soliton solutions of the \((2+1)\)-dimensional BK equation were obtained using this method. These solutions were different from the found solutions in other studies and were presented for the first time in this study. The accuracy of the results was confirmed by putting the obtained solutions back into the original equation. The new solutions of these equations studied could have helped to understand the phenomena in which waves are governed by these equations. In addition, some special values and intervals were given to the results obtained using Wolfram Mathematic 2D and 3D graphical representations of the solutions were made.

The considered method can also be applied to other nonlinear partial differential equations. The most important advantage of this method is that all solutions are obtained from a single algebraic equation. This means that it is sufficient to set up a single algorithm and there is no unnecessary computational overhead.

Author Contribution Statements The authors contributed equally and they read and approved the final manuscript.

Declaration of Competing Interests The authors report that they have no competing interests.

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\title{
ON THE STABILITY ANALYSIS OF A FRACTIONAL ORDER EPIDEMIC MODEL INCLUDING THE GENERAL FORMS OF NONLINEAR INCIDENCE AND TREATMENT FUNCTION
}

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\begin{abstract}
In this paper, we propose to study a SEIR model of fractional order with an incidence and a treatment function. The incidence and treatment functions included in the model are general nonlinear functions that satisfy some meaningful biological hypotheses. Under these hypotheses, it is shown that the disease free equilibrium point of the proposed model is locally and globally asymptotically stable when the reproduction number \(R_{0}\) is smaller than 1. When \(R_{0}>1\), it is established that the endemic equilibrium of the studied system is uniformly asymptotically stable. Finally, some numerical simulations are provided to illustrate the theory.
\end{abstract}

\section*{1. Introduction}

Studying the spread process of infectious diseases has been a very important and popular topic since outbreaks have serious impacts on the economy, daily lives and the future. For finding intervention strategies or treatments and reducing the deaths understanding this mechanism is very important. Mathematical models help us understand the dynamics of epidemiological diseases and talk about the future of the epidemics. Until today, a vast number of mathematical models have been developed for diseases such as rabies, measles, malaria, chickenpox, tuberculosis, cancer, HIV /AIDS and COVID-19 \(1,7,7,-16,18,26,36-38,40,46\). When modeling disease transmission, compartmental models such as SIR (Suscepted-Infected-Recovered), SIS, SEIR and SEIS models are mostly used in the literature (for more detail, see 8\()\). The general idea behind these compartmental models is to divide the total population into compartments and describe the transfer from one compartment to

\footnotetext{
2020 Mathematics Subject Classification. 34D23, 26A33, 92D25, 92D30.
Keywords. Fractional order SEIR model, uniform asymptotic stability, nonlinear incidence function, treatment function.
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another under some meaningful assumptions 21. For example, for an SIR model, the population is divided into three compartments: susceptible individuals \(S\) who are not infected yet, infected individuals \(I\) who are infectious and can spread the disease to the susceptible class and recovered (removed) individuals \(R\) who recovered the infection and already gained the immunity.

In epidemic models, incidence functions are used in the transmission progress between the susceptible population and the infected or exposed population. In literature, there are different incidence rates and using different incidence rates can affect the dynamic behaviour of the system 27,31. The most common incidence rates are the bilinear incidence rate \(\beta S I\) 45,50, where \(\beta\) is the average number of contacts per individual per day, the standard incidence rate \(\beta S I / N, 14,25\) where \(N=S+I+R\), the saturated incidence rate \(\beta S I /\left(1+\alpha_{1} S+\alpha_{2} I\right) 16\) where \(\alpha_{1}, \alpha_{2}\) are positive constants. Even though it is very hard to fit real data values in infectious disease transmission, nonlinearity is inevitable in the incidence rates. These nonlinear incidence rates seem more realistic because they may include saturation effects, heterogeneous mixing populations, environmental factors, media effects or behavioural changes of individuals, etc. 23. In 2005, Korobeinikov and Maini [24 studied the stability properties of infectious disease models with a general, arbitrary nonlinear incidence rate \(f(S, I, N)\) and obtained the global stability by constructing a Lyapunov function under a more specific incidence rate of the form \(g(I) h(S)\). Following this work, Korobeinikov studied the global dynamics of infectious disease models with nonlinear incidence rates in 22 and 23 , respectively. In 2014 , Li et al. 27 considered an SIR epidemic model with a nonlinear arbitrary incidence function \(f(S, I)\) and they improved their model by incorporating a time delay representing the latent period. Recently, in 48, the authors analyzed the stability of a fractional order SEIR model with general incidence rate \(F(S) G(I)\) and in 20, the authors studied the local and global stability of the disease free and endemic equilibrium points of a fractional SIR model with a general incidence function \(f(S, I)\).

Treatment is an important strategy to reduce the number of infected people during outbreaks. Vaccination can be thought one form of treatment to protect against infection before the outbreak 6. One of the most efficient methods of treatment is, of course, hospitalisation, but sometimes the resources of countries are not adequate as in the COVID-19 pandemic. In mathematical models, to reflect and analyse the effect of treatment, scientists have incorporated another treatment class into the SIR model 6, 42, 47, or they used some treatment functions in the models 11,15,28. In 1991, Anderson and May 4 proposed that the treatment function is proportional to the number of infectious people. Following this work, some different modified treatment functions are used in order to reflect the treatment capacity of communities (For further details, see 11,28).

In recent decades, fractional differential equations (FDEs) have gained great attention in the stability analysis of ecological or epidemiological models. Besides

FDEs being generalizations of classical differential equations, using FDEs in epidemiology helps to model in a more realistic way since they reflect history (memory) 16 39. There are different kinds of fractional operators in the literature and these operators may answer distinct real world problems 32. Moreover, because of the memory effect, FDEs will be more suitable for epidemic models. For example, in 2020, Naik et al. 32 , proposed and studied the stability of a FDE system that models COVID-19 pandemic with Atangana-Baleanu or Caputo derivative. They divided the total population into eight groups such as suspected, exposed, symptomatic (infected), asymptomatic, quarantined, treated classes etc. and used the real data from Pakistan. In 2020, Yavuz and Sene 49 studied the stability analysis of a fractional predator-prey model with a harvesting rate. In another paper, Naik et al. 33 has established the global dynamics of a fractional order model for the transmission of HIV epidemic with optimal control in 2020. In 2023, Joshi et al. 16 studied COVID-19 pandemic with Atangana-Baleanu derivative. They used an SIR model with nonlinear Beddington-DeAngelis infection rate and Holling type II treatment rate in their paper. Recently, a vast number of papers containing FDEs for different research areas have been published in the literature \(5,16,17,32,34,41\).

Motivated by the aforementioned works \(20,23,27\), in this paper, we have proposed an SEIR model including FDEs with a general nonlinear incidence function \(f(S, I)\) and a treatment function \(T(I)\) :
\[
\begin{align*}
D_{t}^{\alpha} S(t) & =\lambda-f(S, I)-\mu S \\
D_{t}^{\alpha} E(t) & =f(S, I)-(\beta+\mu+r) E+p T(I) \\
D_{t}^{\alpha} I(t) & =\beta E-(\theta+\mu) I-T(I)  \tag{1}\\
D_{t}^{\alpha} R(t) & =r E+q T(I)-\mu R
\end{align*}
\]
where \(D_{t}^{\alpha} u\) represents Caputo fractional derivative of the function \(u\) with the following initial conditions:
\[
\begin{align*}
& S(0)=S_{0}>0, E(0)=E_{0}>0 \\
& I(0)=I_{0}>0 \text { and } R(0)=R_{0}>0 \tag{2}
\end{align*}
\]

In this model (1), \(S, E, I\) and \(R\) denote the suscepted, exposed, infected, and recovered individuals, respectively. To the best of the author's knowledge, a fractional SEIR model with a general incidence function and treatment function has not been studied yet. We have chosen this model with exposed individuals compartment, as most infectious diseases have an incubation period. Before explaining parameters, we need to emphasize that the originality of this paper comes from the choice of the incidence and treatment functions, \(f(S, I)\) and \(T(I)\) functions in model (1), respectively. These functions have not been determined specifically so that depending on the studied disease, one may choose his/her function according to the spread of the disease and treatment type. Moreover, considering these functions in a general way increases the complexity of the proofs.

In model (1), parameter \(\lambda\) is the recruitment rate which represents the total change in the population and assumed as a positive number, \(\beta\) is the rate at which exposed individuals become infectious (incubation rate), \(\mu\) is the natural death rate, \(\theta\) is the death rate depending on the infection, and \(r\) is the recovery rate of exposed individuals. The function \(T(I)\) represents the general treatment function. In the model, it is assumed that unsuccessfully treated infectious individuals re-enter the exposed compartment proportional to parameter \(p\) and the parameter \(q\) denotes the fraction of infectious individuals whose treatments are successful \((p=1-q)\). The flowchart for the model (1) is given in Figure 1.


Figure 1. The flowchart for the model (1)

The paper is organised as follows: In Section 2, the definition of Caputo fractional derivative is presented and some lemmas are given for the proofs needed for stability analyses. In Section 3, the properties of the incidence function and the treatment function are analysed and the positivity of the solution of system (1) is proved. After that, in Section 4, the equilibrium points of system (1) are determined and the global stability analysis of disease free equilibrium point and the uniform asymptotic stability of endemic equilibrium point are established. Following these theorems, some numerical simulations are carried out to show some examples in Section 5. Finally, we finish this paper with a conclusion part in Section 6.

\section*{2. Preliminaries on the Caputo Fractional Calculus}

We begin by introducing the definition of Caputo fractional derivative.

Definition 1 ( 39\()\). Let \(t_{0}>0, t>t_{0}, \alpha, t_{0}, t \in \mathbb{R}\). The Caputo fractional derivative of order \(\alpha\) of a function \(f \in \mathbb{C}^{n}\) is given by
\[
\begin{equation*}
C t_{0} D_{t}^{\alpha}=\frac{1}{\Gamma(n-\alpha)} \int_{t_{0}}^{t} \frac{f^{(n)}(s)}{(t-s)^{\alpha+1-n}} d s \tag{3}
\end{equation*}
\]
where \(0<n-1<\alpha<n\) and \(\Gamma\) is the Gamma function.
To prove the nonnegativity of the solutions of model (1) we will need the following lemma.

Lemma 1 (Generalized Mean Value Theorem 35). Suppose that \(f \in C[0, a]\) and \(D^{\alpha} f \in(0, a], 0<\alpha \leq 1\). Then one has
\[
\begin{equation*}
f(x)=f(0)+\frac{1}{\Gamma(\alpha)}\left(D^{\alpha} f\right)(\xi) x^{\alpha} \tag{4}
\end{equation*}
\]
with \(0 \leq \xi \leq x, \forall x \in(0, a]\).
Corollary 1 ( 35 ). Suppose that \(f \in \boldsymbol{C}[0, a]\) and \(D^{\alpha} f \in \boldsymbol{C}(0, a]\) for \(0<\alpha \leq 1\). If \(D^{\alpha} f(x) \geq 0 \forall x \in(0, a)\), then \(f(x)\) is non-decreasing for each \(x \in[0, a]\). If \(D^{\alpha} f(x) \leq 0 \forall x \in(0, a)\), then \(f(x)\) is non-increasing for each \(x \in[0, a]\).

Lemma \(2(\boxed{44})\). Let \(x(t) \in \mathbb{R}^{+}\)be a continuous and derivable function. Then, for any time instant \(t \geq t_{0}\)
\[
C t_{0} D_{t}^{\alpha}\left[x(t)-x^{*}-x^{*} \ln \frac{x(t)}{x^{*}}\right] \leq\left(1-\frac{x^{*}}{x(t)}\right) C t_{0} D_{t}^{\alpha} x(t)
\]
\(x^{*} \in \mathbb{R}^{+}, \quad \forall \alpha \in(0,1)\).

\section*{3. Basic Properties of the Model}

In model (1), we assume that the functions \(S, E, I\) and \(R\) and their Caputo fractional derivatives are continuous when \(t>0\).

The general nonlinear incidence function \(f(S, I)\) and the treatment function \(T(I)\) are considered positive, continuously differentiable functions and they satisfy the following hypotheses:

H1) \(f(S, I)>0, f(0, I)=0, f(S, 0)=0\) for all \(S, I>0\).
H2) \(\frac{\partial f(S, I)}{\partial S}>0\) and \(\frac{\partial f(S, I)}{\partial I}>0\) for all \(S, I>0\).
H3) \(\frac{\partial f(S, 0)}{\partial S}=0\) and \(\frac{\partial f(S, 0)}{\partial I}>0\) for all \(S>0\).
H4) \(\frac{f(S, I)}{I} \leq \frac{\partial f(S, 0)}{\partial I}\) for all \(I>0\).
H5) \(T(0)=0\) and \(T^{\prime}(I)>0\) for \(I \geq 0\).
H6) The function \(\frac{T(I)}{I}\) is monotone increasing function, that is,
\[
\frac{T(I)}{I}-T^{\prime}(I) \leq 0 .(\text { See } 11)
\]

These conditions are consistent with biological assumptions and in accordance with literature (see \(11,20,23\) ). For example, for (H1), we can think that there will be no transmission when there are no susceptible or infected people. For (H2), we understand that the incidence function is a monotonically growing function for all \(S, I>0\). In the absence of an infected person, susceptible individuals will become stagnant and transmission will begin to increase in the case of an infected person. In addition, in the absence of an infected person, there is no need for treatment. When there is an increase in the rate of transmission, that is, the number of infected people increases, we need to apply more treatment strategies (H5).

Now, let \(\mathbb{R}_{+}^{4}=\left\{\left(x_{1}, x_{2}, x_{3}, x_{4}\right) \in \mathbb{R}^{4}: x_{i} \geq 0\right.\) for all \(\left.i=1,2,3,4\right\}\). We will prove the existence, uniqueness and positivity of the solutions with the following theorem.

Theorem 1. There exists a unique solution of the model (1) with initial conditions under the hypothesis (H1) is satisfied. Moreover, the solution will remain in \(\mathbb{R}_{+}^{4}\) for all \(t \geq 0\).
Proof. From Theorem 3.1 and Remark 3.2 of 29 , one can see the existence and uniqueness of the solution of system (1) with the initial conditions (2). Now, we will prove that \(\mathbb{R}_{+}^{4}\) is a positively invariant region. For this, let
\[
\begin{gathered}
D_{t}^{\alpha} S(t)_{\mid S=0}=\lambda \geq 0 \\
D_{t}^{\alpha} E(t)_{\mid E=0}=f(S, I)+p T(I) \geq 0 \\
D_{t}^{\alpha} I(t)_{\mid I=0}=\beta E \geq 0
\end{gathered}
\]

We can make a similar discussion as in Theorem 2 of 3 and with the help of Corollary 1, one can observe that the solution will remain in \(\mathbb{R}_{+}^{4}\) for all \(t \geq 0\).

\section*{4. Equilibrium Points and Their Stability}

In this section, we study the stability of the equilibrium points of system (1). Since the right hand sides of the first three equations of model (1) do not include \(R(t)\) we will deal with the first three variables \(S, E\) and \(I\). System (1) has two possible equilibrium points. There is always a disease-free equilibrium point \(E^{0}=\) ( \(S^{0}, 0,0\) ) where \(S^{0}=\frac{\lambda}{\mu}\) provided that (H1) and (H5) is satisfied.

Using the notations in 43, the matrices \(F\) and \(V\) for system (1) are given as follows:
\[
\begin{gathered}
F=\left[\begin{array}{cc}
0 & f_{I}\left(S_{0}, 0\right) \\
0 & 0
\end{array}\right] \\
V=\left[\begin{array}{cc}
(\beta+\mu+r) & -p T^{\prime}(0) \\
-\beta & (\theta+\mu)+T^{\prime}(0)
\end{array}\right] .
\end{gathered}
\]

The basic reproduction number can be found as
\[
\begin{equation*}
R_{0}=\frac{\beta f_{I}\left(S_{0}, 0\right)}{\eta(\theta+\mu)+(\beta q+\mu+r) T^{\prime}(0)} \tag{5}
\end{equation*}
\]
where \(\eta=\beta+\mu+r\).

Lemma 3. Consider a function \(g\) as \(g(I)=\lambda \beta-\eta F(I)+\beta p T(I)\). The equation \(g(I)=0\) has a unique positive solution \(I_{0}\) where \(F(I)=(\theta+\mu) I+T(I)\).
Proof. One can observe that \(g(0)=\lambda \beta>0\) under the condition \(T(0)=0\) and for the positive value \(\frac{\lambda \beta}{\eta(\theta+\mu)}\), we find \(g\left(\frac{\lambda \beta}{\eta(\theta+\mu)}\right)=-(\mu+r+q) T\left(\frac{\lambda \beta}{\eta(\theta+\mu)}\right)<0\) since \(T(I)\) is an increasing function. Moreover,
\[
g^{\prime}(I)=-\eta(\theta+\mu)-(\beta q+\mu+r) T^{\prime}(I)<0
\]

Clearly, the equation \(\lambda \beta-\eta F(I)+\beta p T(I)=0\) has a unique positive solution \(I_{0}\).
Theorem 2. When \(R_{0}>1\), the system (1) has a unique endemic equilibrium point.
Proof. For finding an equilibrium point, let
\(D_{t}^{\alpha} S(t)=0, D_{t}^{\alpha} E(t)=0\) and \(D_{t}^{\alpha} I(t)=0\). Now, we need to solve the following system
\[
\begin{align*}
& \lambda-f(S, I)-\mu S=0 \\
& f(S, I)-\eta E+p T(I)=0  \tag{6}\\
& \beta E-(\theta+\mu) I-T(I)=0
\end{align*}
\]

It is easy to see that \(E^{0}=\left(S^{0}, 0,0\right)\) where \(S^{0}=\frac{\lambda}{\mu}\) is always disease-free equilibrium point of system (1). From system (6), we have
\[
\begin{equation*}
\lambda-\mu S=f(S, I)=\eta E-p T(I) \tag{7}
\end{equation*}
\]
and
\[
\begin{equation*}
E=\frac{F(I)}{\beta} \tag{8}
\end{equation*}
\]
where \(F(I)\) is defined as in Lemma 3. By using the equations (7) and (8) one can obtain \(S\) as in the following form
\[
\begin{equation*}
S=\frac{\lambda}{\mu}-\frac{\eta}{\beta \mu} F(I)+\frac{p}{\mu} T(I) \tag{9}
\end{equation*}
\]

On the other hand, we need to find a positive root for the equation \(f(S, I)=\) \(\eta E-p T(I)\). Let us define the continuous function
\[
\begin{equation*}
H(I)=f\left(\frac{\lambda}{\mu}-\frac{\eta}{\beta \mu} F(I)+\frac{p}{\mu} T(I), I\right)-\frac{\eta}{\beta} F(I)+p T(I) \tag{10}
\end{equation*}
\]

Clearly, \(H(0)=0\) under the hypotheses (H1) and (H5). By using Lemma 3, one can calculate a positive \(I_{0}\) value such that \(\lambda \beta-\eta F\left(I_{0}\right)+\beta p T\left(I_{0}\right)=0\). At this positive \(I_{0}\) value, \(H\left(I_{0}\right)=-\lambda<0\) is obtained.

Moreover, if we look at the derivative of the function \(H(I)\), we can see that
\[
\begin{equation*}
H^{\prime}(I)=f_{I}\left(\frac{\lambda}{\mu}-\frac{\eta}{\beta \mu} F(I)+\frac{p}{\mu} T(I), I\right)-\frac{\eta}{\beta} F^{\prime}(I)+p T^{\prime}(I) \tag{11}
\end{equation*}
\]
and
\[
\begin{equation*}
H^{\prime}(0)=\frac{\left[\eta(\theta+\mu)+(\beta q+\mu+r) T^{\prime}(0)\right]}{\beta}\left(R_{0}-1\right)>0, \tag{12}
\end{equation*}
\]
which implies that there exist some \(I^{*} \in\left(0, I_{0}\right)\) such that \(H\left(I^{*}\right)=0\). Also, at this positive \(I^{*}\) value, \(S^{*}=\frac{1}{\mu}\left(\lambda-\frac{\eta}{\beta} F\left(I^{*}\right)+p T\left(I^{*}\right)\right)>0\) since \(\frac{\mu}{\beta}(\theta+\mu) I^{*}+\) \(\frac{\mu+r}{\beta} T\left(I^{*}\right)+q T\left(I^{*}\right)<\frac{\mu}{\beta}(\theta+\mu) I_{0}+\frac{\mu+r}{\beta} T\left(I_{0}\right)+q T\left(I_{0}\right)=\lambda\) and \(T(I)\) is an increasing function and \(E^{*}=\frac{F\left(I^{*}\right)}{\beta}>0\). This guarantees the existence of a positive endemic equilibruim point \(\Sigma^{*}=\left(S^{*}, E^{*}, I^{*}\right)\). For uniqueness of the endemic equilibrium point, assume that we have another positive equilibrium point \(\tilde{I}\). On the other hand, one can observe \(H^{\prime}\left(I^{*}\right)<0\) which says that at every root of the function \(H(I)\), the function is strictly decreasing. But, this will be a contradiction. This completes the proof. Next, we will prove the local and global asymptotic stability of the disease-free equilibrium point \(E^{0}\).

Theorem 3. The disease-free equilibrium point \(E^{0}=\left(S^{0}, 0,0\right)\) is locally asymptotically stable if \(R_{0}<1\).
Proof. The disease-free equilibrium point \(E^{0}\) is locally asymptotically stable if all of the eigenvalues \(\lambda_{i}\) of the Jacobian matrix evaluated at the equilibrium point \(E^{0}\), that is \(J\left(E^{0}\right)\), satisfy the Matignon's conditions [30, that is,
\[
\begin{equation*}
\left|\arg \left(\lambda_{i}\right)\right|>\frac{\alpha \pi}{2} . \tag{13}
\end{equation*}
\]

The Jacobian matrix \(J\left(E^{0}\right)\) can be evaluated as
\[
J\left(E^{0}\right)=\left[\begin{array}{ccc}
-\mu & 0 & -f_{I}\left(S_{0}, 0\right)  \tag{14}\\
0 & -\eta & f_{I}\left(S_{0}, 0\right)+p T^{\prime}(0) \\
0 & \beta & -\left(\theta+\mu+T^{\prime}(0)\right)
\end{array}\right] .
\]

The eigenvalues of the Jacobian matrix \(J\left(E^{0}\right)\) are \(\lambda_{1}=-\mu\) and \(\lambda_{2,3}=\frac{-(\eta+\sigma) \pm \sqrt{(\eta+\sigma)^{2}-4(\eta \sigma-\gamma)}}{2}\) where \(\sigma=\theta+\mu+T^{\prime}(0)\) and \(\gamma=\) \(\beta\left(f_{I}\left(S_{0}, 0\right)+p T^{\prime}(0)\right)\). It is easy to see that all the eigenvalues are real and if \(\eta \sigma-\gamma>0\) all the eigenvalues \(\lambda_{i}\) will be negative. Hence, \(\left|\arg \left(\lambda_{i}\right)\right|>\frac{\alpha \pi}{2}\). Also, note that the inequality \(\eta \sigma-\gamma>0\) implies that \(R_{0}<1\). Therefore, if \(R_{0}^{2}<1\) the disease-free equilibrium point \(E^{0}\) is locally asymptotically stable and if \(R_{0}>1\) the trivial equilibrium \(E^{0}\) becomes unstable.

Theorem 4. Let
\[
\begin{equation*}
\lim _{I \mapsto 0^{+}} \frac{f\left(S^{0}, I\right)}{f(S, I)}>1 \quad \text { for } \quad S \in\left[0, S^{0}\right) \tag{15}
\end{equation*}
\]

The disease-free equilibrium point \(E^{0}=\left(S^{0}, 0,0\right)\) is globally asymptotically stable if \(R_{0}<1\).

Proof. To establish the global stability of the disease-free equilibrium point, consider the following Lyapunov function
\[
\begin{equation*}
V(t)=\beta E(t)+\eta I(t) \tag{16}
\end{equation*}
\]

Calculating the time fractional order derivative in Caputo sense of both sides of Eq. 16 and using the hypothesis of this theorem and (H4) we get
\[
\begin{align*}
D_{t}^{\alpha} V(t) & =\beta D_{t}^{\alpha} E(t)+\eta D_{t}^{\alpha} I(t) \\
& =\beta I\left(\frac{f(S, I)-\eta E+p T(I)}{I}\right) \\
& +\eta(\beta E-(\theta+\mu) I-T(I) \\
& \leq \beta I\left(f_{I}(S, 0)-\frac{\eta E}{I}+\frac{p T(I)}{I}\right)  \tag{17}\\
& +\eta(\beta E-(\theta+\mu) I-T(I) \\
& =I\left(\beta f_{I}(S, 0)-\eta(\theta+\mu)\right)+(\beta p-\eta) T(I) \\
& \leq I\left(\beta f_{I}\left(S^{0}, 0\right)-\eta(\theta+\mu)\right)+(\beta p-\eta) T(I) \\
& \leq I\left(\eta(\theta+\mu)+(\beta q+\mu+r) T^{\prime}(0)\right)\left(R_{0}-1\right)
\end{align*}
\]

Clearly, if \(R_{0}<1\) then \(D_{t}^{\alpha} V(t)\) is negative. Therefore, the equilibrium point \(E^{0}\) is globally asymptotically stable. This completes the proof.

Theorem 5. Let \(\alpha \in(0,1)\) and \(R_{0}>1\). Then the unique endemic equilibrium point \(\Sigma^{*}=\left(S^{*}, E^{*}, I^{*}\right)\) of system (1) is uniformly asymptotically stable if the following conditions hold
\[
\begin{array}{ll}
\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}<1 & \text { for } \\
\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}>1 & \text { for } \tag{18}
\end{array} \quad S>S^{*},
\]
and
\[
\begin{align*}
& \frac{E}{E^{*}}<\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)} \quad<1 \quad \text { for } \quad E<E^{*}, \\
& \frac{E}{E^{*}}>\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)} \quad>1 \quad \text { for } \quad E>E^{*} \tag{19}
\end{align*}
\]
and
\[
\begin{align*}
& \frac{E}{E^{*}} \leq \frac{T(I)}{T\left(I^{*}\right)}  \tag{20}\\
& \frac{E}{E^{*}} \geq \frac{T(I)}{T\left(I^{*}\right)} \quad \text { for } \quad \frac{E}{E^{*}} \leq \frac{I}{I^{*}} \\
& \geq \frac{I}{I^{*}}
\end{align*}
\]

Proof. We consider the following Lyapunov function
\[
\begin{equation*}
L(t)=L_{1}(S(t))+L_{2}(E(t))+p L_{3}(I(t)) \tag{21}
\end{equation*}
\]
where
\[
\begin{aligned}
& L_{1}(S(t))=S(t)-S^{*}-S^{*} \ln \frac{S(t)}{S^{*}} \\
& L_{2}(E(t))=E(t)-E^{*}-E^{*} \ln \frac{E(t)}{E^{*}} \\
& L_{3}(I(t))=I(t)-I^{*}-I^{*} \ln \frac{I(t)}{I^{*}}
\end{aligned}
\]

Function \(L\) is defined, continuous and positive definite for all \(S(t)>0, E(t)>0\) and \(I(t)>0\). With the help of Lemma 2, we have
\[
\begin{aligned}
C t_{0} D_{t}^{\alpha} L(t) & \leq\left(1-\frac{S^{*}}{S}\right) C t_{0} D_{t}^{\alpha} S(t)+\left(1-\frac{E^{*}}{E}\right) C t_{0} D_{t}^{\alpha} E(t) \\
& +p\left(1-\frac{I^{*}}{I}\right) C t_{0} D_{t}^{\alpha} I(t)
\end{aligned}
\]

Using the equations in system (11) , one has
\[
\begin{aligned}
C t_{0} D_{t}^{\alpha} L(t) & \leq\left(1-\frac{S^{*}}{S}\right)(\lambda-f(S, I)-\mu S) \\
& +\left(1-\frac{E^{*}}{E}\right)(f(S, I)-\eta E+p T(I)) \\
& +p\left(1-\frac{I^{*}}{I}\right)(\beta E-(\theta+\mu) I-T(I))
\end{aligned}
\]

By the equilibrium conditions
\(\lambda=f\left(S^{*}, I^{*}\right)+\mu S^{*}, \eta E^{*}=f\left(S^{*}, I^{*}\right)+p T\left(I^{*}\right)\) and \(\beta E^{*}=(\theta+\mu) I^{*}+T\left(I^{*}\right)\) one can write
\[
\begin{aligned}
C t_{0} D_{t}^{\alpha} L(t) & \leq\left(1-\frac{S^{*}}{S}\right) f\left(S^{*}, I^{*}\right) \\
& -\left(1-\frac{S^{*}}{S}\right) f(S, I)+\mu S^{*}\left(2-\frac{S}{S^{*}}-\frac{S^{*}}{S}\right) \\
& +\left(1-\frac{E^{*}}{E}\right)\left(f(S, I)-f\left(S^{*}, I^{*}\right) \frac{E}{E^{*}}\right) \\
& +\left(1-\frac{E^{*}}{E}\right)\left(-p T\left(I^{*}\right) \frac{E}{E^{*}}+p T(I)\right) \\
& +p\left(1-\frac{I^{*}}{I}\right)\left((\theta+\mu) I^{*}+T\left(I^{*}\right)\right) \frac{E}{E^{*}} \\
& -p\left(1-\frac{I^{*}}{I}\right)(\theta+\mu) I-p\left(1-\frac{I^{*}}{I}\right) T(I)
\end{aligned}
\]

After some computations, we obtain
\[
\begin{aligned}
C t_{0} D_{t}^{\alpha} L(t) & \leq f\left(S^{*}, I^{*}\right)\left(\left(1-\frac{S^{*}}{S}\right)-\left(1-\frac{S^{*}}{S}\right) \frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}\right) \\
& +f\left(S^{*}, I^{*}\right)\left(\left(1-\frac{E^{*}}{E}\right) \frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}\right) \\
& -f\left(S^{*}, I^{*}\right)\left(\left(1-\frac{E^{*}}{E}\right) \frac{E}{E^{*}}\right) \\
& +\mu S^{*}\left(2-\frac{S}{S^{*}}-\frac{S^{*}}{S}\right) \\
& -\left(1-\frac{E^{*}}{E}\right) p T\left(I^{*}\right) \frac{E}{E^{*}}+\left(1-\frac{E^{*}}{E}\right) p T(I) \\
& +p\left(1-\frac{I^{*}}{I}\right)\left((\theta+\mu) I^{*}+T\left(I^{*}\right)\right) \frac{E}{E^{*}} \\
& -p\left(1-\frac{I^{*}}{I}\right)(\theta+\mu) I-p\left(1-\frac{I^{*}}{I}\right) T(I) .
\end{aligned}
\]

Moreover, we get
\[
\begin{aligned}
C t_{0} D_{t}^{\alpha} L(t) & \leq f\left(S^{*}, I^{*}\right)\left(1-\frac{S^{*}}{S}\right)\left(1-\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}\right) \\
& +f\left(S^{*}, I^{*}\right)\left(1-\frac{E^{*}}{E}\right)\left(\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}-\frac{E}{E^{*}}\right) \\
& +\mu S^{*}\left(2-\frac{S}{S^{*}}-\frac{S^{*}}{S}\right) \\
& +p T\left(I^{*}\right)\left(\frac{E^{*}}{E}-\frac{I^{*}}{I}\right)\left(\frac{E}{E^{*}}-\frac{T(I)}{T\left(I^{*}\right)}\right) \\
& +p(\theta+\mu) I^{*}\left(2-\frac{I}{I^{*}}-\frac{I^{*}}{I}\right)
\end{aligned}
\]

By Theorem hypotheses,
\[
\left(1-\frac{S^{*}}{S}\right)\left(1-\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}\right) \leq 0
\]
and
\[
\left(1-\frac{E^{*}}{E}\right)\left(\frac{f(S, I)}{f\left(S^{*}, I^{*}\right)}-\frac{E}{E^{*}}\right) \leq 0
\]
where strict equality holds when \(S=S^{*}, E=E^{*}\). Moreover,
\[
\left(\frac{E^{*}}{E}-\frac{I^{*}}{I}\right)\left(\frac{E}{E^{*}}-\frac{T(I)}{T\left(I^{*}\right)}\right) \leq 0
\]
is satisfied under the assumptions of the theorem. On the other hand,
\[
\left(2-\frac{S}{S^{*}}-\frac{S^{*}}{S}\right) \leq 0
\]
and
\[
\left(2-\frac{I}{I^{*}}-\frac{I^{*}}{I}\right) \leq 0
\]
for all \(S, I>0\), since the arithmetic mean-geometric mean inequality is satisfied. So, by Theorem 3.1 in \(\sqrt[10]{ }\), the positive endemic equilibrium point \(\Sigma^{*}=\left(S^{*}, E^{*}, I^{*}\right)\) of system (1) is uniformly asymptotically stable.

\section*{5. Numerical Simulations}

In this section, we perform some numerical simulations to observe the results obtained in Section 4. We study system (1) for different values of the noninteger order derivative \(\alpha\). For the numerical simulations, the Adams-Bashforth-Moulton scheme 12 is used in Matlab. The following Table 1 shows the parameters that are used in system (1). The assumed values \(p\) and \(q\) have been chosen randomly. The value of \(r\) has taken as small due to the recovery rate of an exposed person should be small. The death rate is again randomly selected that one of the 5 patients died from the disease. The value of the parameter \(w\) is also considered as small and variable in our numerical examples.

Table 1. Parameter values used in numerical simulations
\begin{tabular}{|c|l|l|}
\hline Parameter & Explanation & Value \\
\hline\(\lambda\) & Recruitment rate & 792.857147 \\
\hline\(\mu\) & Natural death rate & \(1 / 7047\) \\
\hline\(\beta\) & Incubation rate & 0.00368 \\
\hline\(r\) & Recovery rate of exposed individuals & 0.01 (Assume) \\
\hline\(p\) & Unsuccessfully treated individuals & 0.8 (Assume) \\
\hline\(q\) & Successfully treated individuals & 0.2 (Assume) \\
\hline\(\theta\) & Death rate depending on the infection & 0.2 (Assume) \\
\hline\(w\) & Transmission coefficient & \(5 \times 10^{-3}\) (Assume) \\
\hline\(r_{1}\) & Treatment function coefficient & 0.511 \\
\hline\(r_{2}\) & Treatment function coefficient & 0.111 \\
\hline\(\gamma\) & Transmission coefficient & 0.119 \\
\hline\(\alpha_{1}\) & Incidence function parameter & 0.01 \\
\hline\(\alpha_{2}\) & Incidence function parameter & 0.01 \\
\hline
\end{tabular}

As a first example, we have chosen bilinear incidence function \(f(S, I)=w S I\) and the treatment function \(T(I)=\frac{r_{1} I^{2}}{1+r_{2} I}\) where \(w, r_{1}\) and \(r_{2}\) are positive parameters. The chosen incidence and treatment functions satisfy our hypotheses. The bilinear


Figure 2. Trajectories for system (11) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=10000\), \(I(0)=0\) and \(R(0)=0\).


Figure 3. Trajectories for system (1) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=10000\), \(I(0)=0\) and \(R(0)=0\).
incidence function is mostly common in literature and it indicates that the rate of transmission increases as the number of infected people increases, that is, as the connection between the susceptible population and the infected population increases. On the other hand, the chosen treatment function is a monotonically increasing function for \(I>0\). On the next page, the readers can see another numerical example with a saturated treatment function \(T(I)=\frac{r_{1} I}{1+r_{2} I}\).

To obtain the value of \(R_{0}\) smaller than one, first we choose the transmission coefficient \(w=5 \times 10^{-6}\) and find \(R_{0}=0.1704<1\). In this case, the diseasefree equilibrium point \(E^{0}=\left(S^{0}, 0,0\right)\) where \(S^{0}=55500\) is globally asymptotically stable which is depicted in Figure 2 and 3.


Figure 4. Trajectories for system (1) with parameters given in Table 1 and the initial conditions \(S(0)=5000, E(0)=20000\), \(I(0)=0\) and \(R(0)=0\).


Figure 5. Trajectories for system (1) with parameters given in Table 1 and the initial conditions \(S(0)=5000, E(0)=20000\), \(I(0)=0\) and \(R(0)=0\).

If we choose \(w=5 \times 10^{-3}\), we calculate \(R_{0}=170.4087>1\). Then, with respect to the calculations in Theorem 2 one can find the approximate value of \(I_{0}\) as 29.368 and the root of the function \(H(I)\) as \(I^{*}=27.0856\). After that, \(S^{*}=5335.5\)


Figure 6. Trajectories for system (1) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=2000\), \(I(0)=0\) and \(R(0)=0\).


Figure 7. Trajectories for system (1) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=2000\), \(I(0)=0\) and \(R(0)=0\).
and \(E^{*}=28455\) can be easily obtained. According to Theorem 5 conditions the equilibrium point \(\Sigma^{*}=\left(S^{*}, E^{*}, I^{*}\right)\) of system (1) is uniformly asymptotically stable which can be seen in Figure 4 and 5

As a second example, we have changed the incidence function as the BeddingtonDeAngelis infection rate, that is, \(f(S, I)=\frac{\gamma S I}{1+\alpha_{1} S+\alpha_{2} I}\). The parameters \(\gamma, \alpha_{1}\) and \(\alpha_{2}\) can be found in Table 1. As in the paper 16, the parameters \(\gamma, \alpha_{1}\) and \(\alpha_{2}\) can be thought as the transmission rate of the disease, a measure of inhibition for the susceptible population, and a measure of inhibition for the infected population,
respectively. In this example, using these parameters, the positive equilibrium point is evaluated as \(\Sigma^{*}=\left(S^{*}, E^{*}, I^{*}\right)=(53221,1278.8,3.26)\) with \(R_{0}=6.1298>1\). The trajectories can be seen in Figure 6 and 7


Figure 8. Trajectories for system (11) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=2000\), \(I(0)=0\) and \(R(0)=0\).


Figure 9. Trajectories for system (11) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=2000\), \(I(0)=0\) and \(R(0)=0\).

Additionally, let us take the saturated treatment function, \(T(I)=\frac{r_{1} I}{1+r_{2} I}\) in model (1) with the Beddington-DeAngelis infection function. In this case, the parameters \(r_{1}\) and \(r_{2}\) represent the treatment rate of disease and the limitation in treatment availability. This treatment function has a horizontal asymptote which
shows the limitations in the capacity of treatment facilities. The numerical simulations can be found in Figure 8 and 9 . Due to limited medical resources, after a while there will be almost no susceptible individuals left, and almost all people can be exposed, infected, or recovered from the disease. (Remember that we have a unique positive equilibrium point and in Figure 8 case (a), \(S(t)\) values are approximately 115 not zero.) If we increase the \(r_{2}\) parameter, i.e., the limitations of medical resources, we can see that the number of infected people increases as we expected in Figure 10


Figure 10. Trajectories for system (1) with parameters given in Table 1 and the initial conditions \(S(0)=50000, E(0)=2000\), \(I(0)=0\) and \(R(0)=0\).

Following this second example, we also wondered the effect of the treatment success. As we expect, if the rate of successfully treated individuals \(q\) increase, we observe a decrease in the number of the exposed and infected individuals in Figure (11)

Finally, we need to mention our observation about the parameter \(\alpha\). If we look carefully at all the figures, we will see that the system will be stable over a longer period of time if the parameter \(\alpha\) decreases.

\section*{6. Concluding Remarks}

In this paper, we have introduced a fractional order SEIR model with a general incidence function \(f(S, I)\) and a general treatment function \(T(I)\). By analysing the equilibrium points of system (1), we have shown that the disease free equilibrium point is locally asymptotically stable if the basic reproduction number \(R_{0}<1\) and globally asymptotically stable if the inequality 15 is satisfied when \(R_{0}<\) 1. We have also constructed a Lyapunov function and found that the endemic


Figure 11. Trajectories for system (1) with parameters given in Table 1 and changing parameters \(p\) and \(q\) and the initial conditions \(S(0)=50000, E(0)=2000, I(0)=0\) and \(R(0)=0\).
equilibrium point is uniformly asymptotically stable under the conditions (18)(20). Moreover, in our model, based on the treatment model in 43, unsuccessfully treated infectious individuals re-enter the exposed compartment proportional to parameter \(p\). Changing these parameters \(p\) and \(q\), we also have a chance to compare treatment success.

To the best of the author's knowledge, a fractional SEIR model with a general incidence function and treatment function has not been studied yet. In 2017, Elkhaiar et al. 11 studied the stability analysis of an ordinary differential equation system of the SEIR model with treatment. In that paper, they used a general incidence function and a general treatment function \(T^{*}(I)\). In our paper, if the parameters are changed as \(\mu=d, p=0, q=1, r=0, \beta=\sigma, \theta=0\) and \(T(I)=\gamma I+T^{*}(I)\), then system (1) becomes to the SEIR model studied in 11 when \(\alpha=1\). Another paper including a stability analysis of a fractional order SEIR model is published in 2020 by Yang et. al. 48. In our system, if the parameters are chosen as \(\lambda=\Lambda^{\alpha}\), \(\mu=d^{\alpha}, r=0, \beta=\sigma^{\alpha}, \theta=0\) and \(q T(I)=\gamma^{\alpha} I, f(S, I)=\beta^{\alpha} F(S) G(I)\) then system (1) turns into the system that studied in 48. In 2018, analysis of a fractional order SEIR model with treatment is established by Almeida 2. If the parameters in our system (1) is chosen as \(\lambda=b N\) ( \(N\) is assumed as fixed population size), \(\mu=b\), \(r=0, \beta=\sigma, p=0, q=1, \theta=0\) and \(f(S, I)=\frac{\beta I S}{N}\) and \(T(I)=(\mu+q) I(\mu\) and \(q\) are the parameters used in 22) then again system (1) turns into that system used in 2 . Finally, the last example studied in 13 is related to outbreaks of influenza \(\mathrm{A}(\mathrm{H} 1 \mathrm{~N} 1)\). In \(\sqrt[13]{ }\), the authors proposed a fractional order SEIR model to explain and understand the outbreaks of ifluenza \(\mathrm{A}(\mathrm{H} 1 \mathrm{~N} 1)\). They used real data values and tested and simulated these data values for their model and chose the best fitted order \(\alpha\) of fractional differentiation. In system (1), if the parameters are changed
as \(\lambda=\mu^{\alpha}, \mu=\mu^{\alpha}, \beta=\Omega^{\alpha}, p=q=\theta=0, q=1, f(S, I)=\beta^{\alpha} S I\) and \(T(I)=\rho^{\alpha} I\), then one can obtain the system studied in this paper.

As a conclusion, the proposed model in this paper is rather general and as pointed out in 13, in the applications, one can choose the best fitted order of \(\alpha\) depending on the real data values. If we look carefully at all the figures, we will see that the system will be stable over a longer period of time if the parameter \(\alpha\) decreases.

Declaration of Competing Interests The author has no competing interests to declare.

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[^7]:    2020 Mathematics Subject Classification. 47H10, 54H25.
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