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[3] P. Erdős, On the distribution of the roots of orthogonal polynomials, in: G. Alexits, S. B. Steckhin (Eds.), Proceedings of a Conference on Constructive Theory of Functions, Akademiai Kiado, Budapest, 1972, pp. 145-150.
[4] D. Allen, Relations between the local and global structure of finite semigroups, Ph. D. Thesis, University of California, Berkeley, 1968.

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# BANACH FIXED POINT THEOREM ON ORTHOGONAL CONE METRIC SPACES 

Zeinab Eivazi Damirchi Darsi Olia, Madjid Eshaghi Gordji and Davood Ebrahimi Bagha

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Abstract. In this paper, we introduce new concept of orthogonal cone metric spaces. We establish new versions of fixed point theorems in incomplete orthogonal cone metric spaces. As an application, we show the existence and uniqueness of solution of the periodic boundry value problem.
Keywords: Orthogonal set; Fixed point; Orthogonal cone metric space; Differential equation; Solution.

## 1. Introduction

The concept of standard metric spaces is a fundamental tool in topology, functional analysis and nonlinear analysis. This structure has attracted a considerable attention from mathematicians because of the development of the fixed point theory in standard metric spaces.
In recent years, several generalizations of standard metric spaces have appeared. Huang and Zhang [8] have introduced the concept of cone metric space by replacing the set of real numbers by an ordered Banach space and proved many fixed point theorems of contractive type mappings in cone metric space. In 2010, W.S.Du [2] has shown that many results in fixed point theory on cone metric spaces are equivalent to ordinary metric spaces. Subsequently, many authors in $[2,7,9]$ have generalized the results of Huang and Zhang [8].

Huang and Zhang [8] considered the concept of cone metric spaces as follows:

Definition 1.1. [8] Let $E$ always be a real Banach space and $P$ a subset of $E . P$ is called a cone if and only if:

1. $P$ is closed, nonempty, and $P \neq\{0\}$;

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2. $a, b \in \mathbb{R}, a, b \geq 0, x, y \in P$ then $a x+b y \in P$;
3. $x \in P$ and $-x \in P$ then $x=0$.

Given a cone $P \subset E$, we define a partial ordering $\leqslant$ with respect to $P$ by $x \leqslant y$ if and only if $y-x \in P$. We shall write $x<y$ to indicate that $x \leqslant y$ but $x \neq y$, while $x \ll y$ will stand for $y-x \in \operatorname{int} P$, int $P$ denotes the interior of $P$. The cone $P$ is called normal if there is a number $K>0$ such that for all $x, y \in E$,

$$
0 \leq x \leq y \Longrightarrow\|x\| \leq K\|y\|
$$

The least positive number satisfying above is called the normal constant of $P$.
The cone $P$ is called regular if every increasing sequence which is bounded from above is convergent. That is, if $\left\{x_{n}\right\}$ is sequence such that

$$
x_{1} \leq x_{2} \leq \quad \cdots \quad \leq x_{n} \leq \cdots \leq y
$$

for some $y \in E$, then there exists $x \in E$ such that $\left\|x_{n}-x\right\| \rightarrow 0$ as $n \longrightarrow \infty$. Equivalently the cone $P$ is regular if and only if every decreasing sequence which is bounded from below is convergent. It is well known that a regular cone is a normal cone.

Definition 1.2. [8] Let $X$ be a nonempty set. Suppose the mapping $d: X \times X \rightarrow E$ satisfies

- (d1) $0<d(x, y)$ for all $x, y \in X$ and $d(x, y)=0$ if and only if $x=y$;
- (d2) $d(x, y)=d(y, x)$ for all $x, y \in X$;
- (d3) $d(x, y) \leq d(x, z)+d(y, z)$ for all $x, y, z \in X$.

Then $d$ is called a cone metric on $X$, and $(X, d)$ is called a cone metric space.
It is obvious that cone metric spaces generalize metric spaces.
Example 1.1. Let $E=\mathbb{R}^{2}, P=\{(x, y) \in E \mid x, y \geq 0\} \subset \mathbb{R}^{2}, X=\mathbb{R}$ and $d: X \times X \rightarrow E$ such that $d(x, y)=(|x-y|, \alpha|x-y|)$, where $\alpha \geq 0$ is a constant. Then $(X, d)$ is a cone metic space.

Definition 1.3. [8] Let $(X, d)$ be a cone metric space. Let $\left\{x_{n}\right\}$ be a sequence in $X$ and $x \in X$. If for every $c \in E$ with $0 \ll c$ there is $N$ such that for all $n>N$, $d\left(x_{n}, x\right) \ll c$, then $\left\{x_{n}\right\}$ is said to be convergent and $\left\{x_{n}\right\}$ converges to $x$ and $x$ is the limit of $\left\{x_{n}\right\}$. We denote this by

$$
\lim _{n \rightarrow \infty} x_{n}=x \quad \text { or } x_{n} \rightarrow x \quad(n \rightarrow \infty)
$$

Definition 1.4. [8] Let $(X, d)$ be a cone metric space, $\left\{x_{n}\right\}$ be a sequence in $X$. If for any $c \in E$ with $0 \ll c$, there is $N$ such that for all $n, m>N, d\left(x_{n}, x_{m}\right) \ll c$, then $\left\{x_{n}\right\}$ is called a Cauchy sequence in $X$.

Definition 1.5. [8] Let $(X, d)$ be a cone metric space if every Cauchy sequence is convergent in $X$, then $X$ is called a complete cone metric space.

Huang and Zhang [8] also proved the following fixed point theorem in cone metric spaces.

Theorem 1.1. [8] Let $(X, d)$ be a complete cone metric space, $P$ be a normal cone with normal constant $K$. Suppose the mapping $T: X \rightarrow X$ satisfies the contractive condition

$$
d(T x, T y) \leq k d(x, y)
$$

for all $x, y \in X$ where $k \in[0,1)$ is a constant. Then $T$ has a unique fixed point in $X$ and for any $x \in X$, an iterative sequence $\left\{T^{n} x\right\}$ converges to the fixed point.

Eshaghi and et.al. [3] introduced the notion of orthogonal sets as follows (also see $[11,1,4,5,6,10])$ :

Definition 1.6. [3] Let $X \neq \phi$ and $\perp \subseteq X \times X$ be a binary relation. If $\perp$ satisfies the following condition

$$
\exists x_{0} ; \quad\left(\left(\forall y ; y \perp x_{0}\right) \text { or } \quad\left(\forall y ; x_{0} \perp y\right)\right)
$$

it is called an orthogonal set (briefly $O$-set). We denote this $O$-set by $(X, \perp)$.
Definition 1.7. Let $(X, \perp)$ be an $O$-set. A sequence $\left\{x_{n}\right\}_{n \in \mathbb{N}}$ is called an orthogonal sequence (briefly $O$-sequence) if

$$
\left(\left(\forall n ; x_{n} \perp x_{n+1}\right) \text { or } \quad\left(\forall n ; x_{n+1} \perp x_{n}\right)\right) .
$$

for more information refer to [3].
Definition 1.8. [3] Let $(X, d, \perp)$ be an orthogonal metric space $((X, \perp)$ is an $O$ set and $(X, d)$ is a metric space). The space $X$ is orthogonally complete (briefly O-complete) if every Cauchy O-sequence is convergent.

It is easy to see that every complete metric space is O-complete and the converse is not true (see [3]).

Definition 1.9. [3] Let $(X, d, \perp)$ be an orthogonal metric space and $0<k<1$.

1. A mapping $f: X \rightarrow X$ is said to be orthogonal contractive ( $\perp$-contractive) mapping with Lipschitz constant $k$ if

$$
d(f x, f y) \leq k d(x, y) \quad \text { if } \quad x \perp y
$$

2. A mapping $f: X \rightarrow X$ is called an orthogonal preserving ( $\perp-$ preserving) mapping if $x \perp y$ then $f(x) \perp f(y)$.
3. A mapping $f: X \rightarrow X$ is an orthogonal continuous ( $\perp$-continuous) mapping in $a \in X$ if for each $O$-sequence $\left\{a_{n}\right\}_{n \in \mathbb{N}}$ in $X$ if $a_{n} \rightarrow a$ then $f\left(a_{n}\right) \rightarrow f(a)$. Also $f$ is $\perp$-continuous on $X$ if $f$ is $\perp$-continuous in each $a \in X$.

They also, proved the following theorem which can be considered as a real extension of Banach fixed point theorem $[11,1,3,4,5,6,10]$.

Theorem 1.2. [3] Let $(X, d, \perp)$ be an $O$-complete metric space (not necessarily complete metric space). Let $f: X \rightarrow X$ be $\perp$-continuous, $\perp$-contraction (with Lipschitz constant $k$ ) and $\perp$-preserving, then $f$ has a unique fixed point $x^{*}$ in $X$. Also, $f$ is a Picard operator, that is, $\lim f^{n}(x)=x^{*}$ for all $x \in X$.

Let us consider the following periodic boundry value problem

$$
\left\{\begin{array}{l}
u^{\prime}(t)=f(t, u(t))  \tag{1.1}\\
u\left(t_{0}\right)=u(T)
\end{array}\right.
$$

where $t \in I=[0, T], T>0$ and $f: I \times \mathbb{R} \rightarrow \mathbb{R}$ is a continuous function. Suppose that there exists $\beta>0, \mu>0$ with $\mu<\beta$ such that for $x, y \in \mathbb{R}$ we have

$$
\begin{equation*}
0 \leq|[f(t, y)+\beta y]-[f(t, x)+\beta x]| \leq \mu|y-x| \tag{1.2}
\end{equation*}
$$

Inspired and motivated by the above results, we introduce new concept of orthogonal cone metric space. In such space, we establish new versions of fixed point theorems. As an application, we show the existence and uniqueness of solution of the periodic boundry value problem 1.1.

## 2. Main Results

In this section, we shall introduce a new definitions to prove the main results. We begin with the following definition. In the following part, we shall suppose $E$ is a Banach space, $P$ is a cone in $E$ with $\operatorname{int} P \neq \phi$ and $\leq$ is partial ordering with respect to $P$.

Definition 2.1. Let $(X, \perp)$ be a nonempty orthogonal set. Suppose the mapping $d: X \times X \rightarrow E$ satisfies

- (d1) $0<d(x, y)$ for all $x, y \in X$ and $d(x, y)=0$ if and only if $x=y$;
- (d2) $d(x, y)=d(y, x)$ for all $x, y \in X$;
- (d3) $d(x, y) \leq d(x, z)+d(y, z)$ for all $x, y, z \in X$.

Then $d$ is called a cone metric on $(X, \perp)$ and $(X, d, \perp)$ is called an orthogonal cone metric space.

We have the concept of orthogonal complete cone metric space as follows:
Definition 2.2. Let $(X, d, \perp)$ be an orthogonal cone metric space, if every Cauchy $O$-sequence is convergent in $X$, then $X$ is called an orthogonal complete cone metric space.

It is easy to see that every complete cone metric space is O-complete and the converse is not true. In the next example, X is O -complete cone metric space and it is not complete.

Example 2.1. Let $E=\mathbb{R}, P=[0, \infty)$ and $X=[0,1)$. Suppose $x \perp y$ if $x \leq y . \quad(X, \perp)$ is an O-set. Clearly, $X$ with metric $d: X \times X \rightarrow E$ such that $d(x, y)=|x-y|$ is not complete cone metric space but it is O-complete cone metric space. Because if $\left\{x_{k}\right\}$ is an arbitrary Cauchy O-sequence in $X$, then there exists a subsequence $\left\{x_{k_{n}}\right\}$ of $\left\{x_{k}\right\}$ for which $\left\{x_{k_{n}}\right\} \leq \frac{1}{2}$ for all $n$. It follows that $\left\{x_{k_{n}}\right\}$ converges to a $x \in\left[0, \frac{1}{2}\right] \subset X$. On the other hand, we know that every Cauchy sequence with a convergent subsequence is convergent. It follows that $\left\{x_{k}\right\}$ is convergent.

In the following example, we shall prove a theorem that can be considered as the main result of our paper.

Theorem 2.1. Let $(X, d, \perp)$ be an orthogonal complete cone metric space (not necessarily complete cone metric space), $P$ be a normal cone with normal constant $K$. Suppose the mapping $T: X \rightarrow X$ is $\perp$-preserving, $\perp$-continuous and $\perp$-contraction Lipschitz constant $k \in[0,1)$. Then $T$ has a unique fixed point in $X$. In addition $T$ is a picard operator.

Proof. By definition of orthogonality, there exists $x_{0} \in X$ such that

$$
\left(\forall x \in X, x \perp x_{0}\right) \quad \text { or } \quad\left(\forall x \in X, x_{0} \perp x\right) .
$$

It follows that $x_{0} \perp T\left(x_{0}\right)$ or $T\left(x_{0}\right) \perp x_{0}$. Let

$$
x_{1}:=T\left(x_{0}\right), x_{2}:=T\left(x_{1}\right)=T^{2}\left(x_{0}\right), \cdots, x_{n+1}:=T x_{n}=T^{n}\left(x_{0}\right) .
$$

We have

$$
\begin{aligned}
d\left(x_{n+1}, x_{n}\right) & =d\left(T x_{n}, T x_{n-1}\right) \\
& \leq k d\left(x_{n}, x_{n-1}\right) \\
& \leq k^{2} d\left(x_{n-1}, x_{n-2}\right) \\
& \vdots \\
& \leq k^{n} d\left(x_{1}, x_{0}\right) .
\end{aligned}
$$

So for $n>m$,

$$
\begin{aligned}
d\left(x_{n}, x_{m}\right) & \leq d\left(x_{n}, x_{n-1}\right)+d\left(x_{n-1}, x_{n-2}\right)+\cdots+d\left(x_{m-1}, x_{m}\right) \\
& \leq\left(k^{n-1}+k^{n-2}+\cdots+k^{m}\right) d\left(x_{1}, x_{0}\right) \\
& \leq \frac{k^{m}}{1-k} d\left(x_{1}, x_{0}\right)
\end{aligned}
$$

We get $\left\|d\left(x_{n}, x_{m}\right)\right\| \leq \frac{k^{m}}{1-k} K\left\|d\left(x_{1}, x_{0}\right)\right\|$. This implies that $d\left(x_{n}, x_{m}\right) \rightarrow 0(n, m \rightarrow$ $\infty)$. Hence the O-sequence $\left\{x_{n}\right\}$ is Cauchy. By completeness of $X$, there exists $x^{*}$ in $X$ such that $x_{n} \rightarrow x^{*}(n \rightarrow \infty)$. On the other hand, $T$ is $\perp$-continuous and hence $T x_{n} \rightarrow T x^{*}$ as $n$ tends to infinity and $T\left(x^{*}\right)=\lim _{n \rightarrow \infty} T\left(x_{n}\right)=\lim _{n \rightarrow \infty} x_{n+1}=x^{*}$. Therefore $x^{*}$ is a fixed point of $T$.

To prove the uniqueness of the fixed point, let $y^{*} \in X$ be a fixed point of $T$. Then we have $T^{n}\left(y^{*}\right)=y^{*}$ for all $n \in \mathbb{N}$. By our choice of $x_{0}$ in the first part of the proof, we have

$$
x_{0} \perp y^{*} \quad \text { or } \quad y^{*} \perp x_{0} .
$$

Since $T$ is $\perp$-preserving, we have

$$
T^{n}\left(x_{0}\right) \perp T^{n}\left(y^{*}\right) \quad \text { or } \quad T^{n}\left(y^{*}\right) \perp T^{n}\left(x_{0}\right)
$$

for all $n \in \mathbb{N}$. On the other hand, $T$ is $\perp$-contraction, then we have for all $n \in \mathbb{N}$,

$$
\begin{aligned}
d\left(x^{*}, y^{*}\right) & =d\left(T^{n}\left(x^{*}\right), T^{n}\left(y^{*}\right)\right) \\
& \leq d\left(T^{n}\left(x^{*}\right), T^{n}\left(x_{0}\right)\right)+d\left(T^{n}\left(x_{0}\right), T^{n}\left(y^{*}\right)\right) \\
& \leq k^{n}\left[d\left(x^{*}, x_{0}\right)+d\left(x_{0}, y^{*}\right)\right] .
\end{aligned}
$$

Also we have

$$
\left\|d\left(x^{*}, y^{*}\right)\right\| \leq K\left(k^{n}\left[\left\|d\left(x^{*}, x_{0}\right)\right\|+\left\|d\left(x_{0}, y^{*}\right)\right\|\right]\right)
$$

As $n$ goes to infinity, we get $x^{*}=y^{*}$.
Finally, we show that $T$ is a Picard operator. Let $x \in X$ be arbitrary. Similarly, then

$$
\left[x_{0} \perp x^{*} \quad \text { and } \quad x_{0} \perp x\right] \text { or }\left[x^{*} \perp x_{0} \text { and } x \perp x_{0}\right] .
$$

Now, since $T$ is $\perp$-preserving, then

$$
\left[T^{n}\left(x_{0}\right) \perp T^{n}\left(x^{*}\right) \text { and } T^{n}\left(x_{0}\right) \perp T(x)\right] \text { or }\left[T^{n}\left(x^{*}\right) \perp T^{n}\left(x_{0}\right) \text { and } T(x) \perp T^{n}\left(x_{0}\right)\right]
$$

for all $n \in \mathbb{N}$. Hence for all $n \in \mathbb{N}$, we get

$$
d\left(T^{n}(x), T^{n}\left(x_{0}\right)\right) \leq k d\left(T^{n-1}(x), T^{n-1}\left(x_{0}\right)\right) \leq \cdots \leq k^{n} d\left(x, x_{0}\right)
$$

Letting $n \rightarrow \infty$ we have $\lim _{n \rightarrow \infty} T^{n}(x)=x^{*}$. This completes the proof.
Here, we obtain another fixed point theorem by replacing $\perp$-contractive condition by another slightly modified condition.

Theorem 2.2. Let $(X, d, \perp)$ be an orthogonal complete cone metric space, $P$ be a normal cone with normal constant $K$. Let $T: X \rightarrow X$ be $\perp$-preserving, $\perp$ continuous mapping satisfying the following $\perp$-contractive condition

$$
d(T x, T y) \leq a(d(x, y))+b[d(x, T x)+d(y, T y)]+c[d(x, T y)+d(y, T x)]
$$

for $x, y \in X$ with $x \perp y$ and the constants $a, b, c \in[0,1)$ and $a+b+c<1$. Then $T$ has a unique fixed point in $X$.

Proof. By definition of orthogonality, there exists $x_{0} \in X$ such that

$$
\left(\forall x \in X, x \perp x_{0}\right) \text { or } \quad\left(\forall x \in X, x_{0} \perp x\right) .
$$

It follows that $x_{0} \perp T\left(x_{0}\right)$ or $T\left(x_{0}\right) \perp x_{0}$. Let

$$
x_{1}:=T\left(x_{0}\right), x_{2}:=T\left(x_{1}\right)=T^{2}\left(x_{0}\right), \cdots, x_{n+1}:=T x_{n}=T^{n}\left(x_{0}\right) .
$$

We have

$$
\begin{aligned}
& d\left(x_{n+1}, x_{n}\right)=d\left(T x_{n}, T x_{n-1}\right) \\
& \leq a\left(d\left(x_{n}, x_{n-1}\right)\right)+b\left[d\left(x_{n}, T x_{n}\right)+d\left(x_{n-1}, T x_{n-1}\right)\right]+c\left[d\left(x_{n}, T x_{n-1}\right)+d\left(x_{n-1}, T x_{n}\right)\right] \\
& \leq a\left(d\left(x_{n}, x_{n-1}\right)\right)+b\left[d\left(x_{n}, x_{n+1}\right)+d\left(x_{n-1}, x_{n}\right)\right]+c\left[d\left(x_{n}, x_{n}\right)+d\left(x_{n-1}, x_{n+1}\right)\right] \\
& \leq a\left(d\left(x_{n}, x_{n-1}\right)\right)+b\left[d\left(x_{n}, x_{n+1}\right)+d\left(x_{n-1}, x_{n}\right)\right]+c\left[d\left(x_{n-1}, x_{n}\right)+d\left(x_{n}, x_{n+1}\right)\right] .
\end{aligned}
$$

Therefore,

$$
d\left(x_{n+1}, x_{n}\right)(1-b-c)=d\left(x_{n}, x_{n-1}\right)(a+b+c)
$$

and we get

$$
d\left(x_{n+1}, x_{n}\right) \leq d\left(x_{n}, x_{n-1}\right) \frac{a+b+c}{1-b-c}
$$

Substituting $k=\frac{a+b+c}{1-b-c}$ and as $0 \leq k<1$ we have

$$
d\left(x_{n+1}, x_{n}\right) \leq k d\left(x_{n}, x_{n-1}\right) \leq \cdots \leq k^{n} d\left(x_{1}, x_{0}\right)
$$

For any $m \geq 1, p \geq 1$, it follows that

$$
\begin{aligned}
d\left(x_{m+p}, x_{m}\right) & \leq d\left(x_{m+p}, x_{m+p-1}\right)+d\left(x_{m+p-1}, x_{m}\right) \\
& \leq d\left(x_{m+p}, x_{m+p-1}\right)+d\left(x_{m+p-1}, x_{m+p-2}\right)+d\left(x_{m+p-2}, x_{m}\right) \\
& \leq d\left(x_{m+p}, x_{m+p-1}\right)+d\left(x_{m+p-1}, x_{m+p-2}\right)+d\left(x_{m+p-2}, x_{m+p-3}\right) \\
& +\cdots+d\left(x_{m+2}, x_{m+1}\right)+d\left(x_{m+1}, x_{m}\right) \\
& \leq k^{m+p-1} d\left(x_{1}, x_{0}\right)+k^{m+p-2} d\left(x_{1}, x_{0}\right)+k^{m+p-3} d\left(x_{1}, x_{0}\right) \\
& +\cdots+k^{m} d\left(x_{1}, x_{0}\right) \\
& \leq\left(k^{m+p-1}+k^{m+p-2}+k^{m+p-3}+\cdots+k^{m}\right) d\left(x_{1}, x_{0}\right) \\
& \leq \frac{k^{m}}{1-k} d\left(x_{1}, x_{0}\right) .
\end{aligned}
$$

So we have

$$
\left\|d\left(x_{m+p}, x_{m}\right)\right\| \leq K \frac{k^{m}}{1-k}\left\|d\left(x_{1}, x_{0}\right)\right\|
$$

Letting $m \longrightarrow \infty$ we conclude that $\left\{x_{n}\right\}$ is a Cauchy O-sequence. Since $(X, d)$ is a complete orthogonal cone metric space, there exists $x^{*} \in X$ such that $x_{n} \rightarrow x^{*}$ as $n \longrightarrow \infty$.

Next, we claim that $x^{*}$ is a fixed point of $T$.

$$
\begin{aligned}
d\left(T x^{*}, x^{*}\right) & \leq d\left(T x^{*}, T x_{n}\right)+d\left(T x_{n}, x^{*}\right) \\
& \leq d\left(T x^{*}, T x_{n}\right)+d\left(x_{n+1}, x^{*}\right)
\end{aligned}
$$

and we have

$$
\begin{aligned}
d\left(T x^{*}, x^{*}\right) & \leq a\left(d\left(x^{*}, x_{n}\right)\right)+b\left[d\left(x^{*}, T x^{*}\right)+d\left(x_{n}, T x_{n}\right)\right] \\
& +c\left[d\left(x^{*}, T x_{n}\right)+d\left(x_{n}, T x^{*}\right)\right]+d\left(x_{n+1}, x^{*}\right) \\
& \leq a\left(\left(d\left(x^{*}, x_{n}\right)\right)+b\left[d\left(x^{*}, T x^{*}\right)+d\left(x_{n}, x_{n+1}\right)\right]\right. \\
& +c\left[d\left(x^{*}, x_{n+1}\right)+d\left(x_{n}, T x^{*}\right)\right]+d\left(x_{n+1}, x^{*}\right) \\
& \leq a\left(\left(d\left(x^{*}, x_{n}\right)\right)+b\left[d\left(x^{*}, T x^{*}\right)+d\left(x_{n}, x^{*}\right)+d\left(x^{*}, x_{n+1}\right)\right]\right. \\
& +c\left[d\left(x^{*}, x_{n+1}\right)+d\left(x_{n}, x^{*}\right)+d\left(x^{*}, T x^{*}\right)\right]+d\left(x_{n+1}, x^{*}\right) .
\end{aligned}
$$

So

$$
d\left(T x^{*}, x^{*}\right)(1-b-c) \leq d\left(x^{*}, x_{n}\right)(a+b+c)+d\left(x^{*}, x_{n+1}\right)(1+b+c),
$$

and

$$
d\left(T x^{*}, x^{*}\right) \leq \frac{d\left(x^{*}, x_{n}\right)(a+b+c)+d\left(x^{*}, x_{n+1}\right)(1+b+c)}{(1-b-c)} .
$$

Therefore

$$
\left\|d\left(T x^{*}, x^{*}\right)\right\| \leq K\left(\frac{(a+b+c)}{(1-b-c)}\left\|d\left(x^{*}, x_{n}\right)\right\|+\frac{(1+b+c)}{(1-b-c)}\left\|d\left(x^{*}, x_{n+1}\right)\right\|\right)
$$

Letting $n \longrightarrow \infty$, we have $T x^{*}=x^{*}$.
Finally, we need to prove that the fixed point is unique.
If there is another fixed point $y^{*}$, then

$$
\begin{aligned}
d\left(x^{*}, y^{*}\right)= & d\left(T x^{*}, T y^{*}\right) \\
\leq & a\left(d\left(x^{*}, y^{*}\right)\right)+b\left[d\left(x^{*}, T x^{*}\right)+d\left(y^{*}, T y^{*}\right)\right]+c\left[d\left(x^{*}, T y^{*}\right)+d\left(y^{*}, T x^{*}\right)\right] \\
\leq & a\left(d\left(x^{*}, y^{*}\right)\right)+b\left[d\left(x^{*}, T x^{*}\right)+d\left(y^{*}, T y^{*}\right)\right] \\
+ & c\left[d\left(x^{*}, T x^{*}\right)+d\left(T x^{*}, T y^{*}\right)+d\left(y^{*}, T y^{*}\right)+d\left(T y^{*}, T x^{*}\right)\right] \\
= & (a+2 c) d\left(x^{*}, y^{*}\right) \\
& (1-a-2 c) d\left(x^{*}, y^{*}\right) \leq 0
\end{aligned}
$$

this implies that

$$
\left\|d\left(x^{*}, y^{*}\right)\right\|=0
$$

Hence $x^{*}=y^{*}$. Therefore the proof is completed.

## 3. Application in differential equations

In this section, we apply results in the previous section to show the existence and uniqueness of solution of the following periodic boundary value problem 1.1 where $t \in I=[0, T], T>0$ and $f: I \times \mathbb{R} \rightarrow \mathbb{R}$ is a continuous function. Let
$X=\{u \in C(I, \mathbb{R}) ; u(t)>1 \quad\{$ for almost every $\} \quad t \in I\}$. Consider the Banach space $E=\mathbb{R}$ and $P=[0, \infty)$. Define partial ordering $\leq$ with respect to $P$ by $a \leq b$ if and only if $b-a \in P$.
Suppose the mapping $d: X \times X \rightarrow E$ by

$$
d(x, y)=\sup _{t \in I}|x(t)-y(t)|,
$$

for $x, y \in X$.
Suppose that there exists $\beta>0, \mu>0$ with $\mu<\beta$ such that for $x, y \in \mathbb{R}$ we have 1.2.

Theorem 3.1. Under above conditions, for all $T>0$ the differential equation 1.1 has a unique solution.

Proof. The problem can be written in integral equation as

$$
x(t)=\int_{0}^{T} G(t, s)[f(s, x(s))+\beta x(s)] d s
$$

where

$$
G(t, s)= \begin{cases}\frac{e^{\beta(T+s-t)}}{e^{\beta T}-1}, & 0 \leq s \leq t \leq T  \tag{3.1}\\ \frac{e^{\beta(s-t)}}{e^{\beta T}-1}, & 0 \leq t \leq s \leq T\end{cases}
$$

Define the following orthogonality relation $\perp$ in $X$ :

$$
x \perp y \quad \text { if } \quad x(t) y(t) \geq y(t)
$$

for almost every $t \in I$. It's easy to see that $(X, d, \perp)$ is a cone metric space. Since every $x$ is a continuous function over a closed and bounded subset of the Euclidean space, this supremum is actually attained in $(X, d, \perp)$. Hence $(X, d, \perp)$ is complete.

Now, we define $A:(X, d, \perp) \rightarrow(X, d, \perp)$ as follows:

$$
(A x)(t)=\int_{0}^{T} G(t, s)[f(s, x(s))+\beta x(s)] d s
$$

for all $t \in I$.
Note that the fixed points of $A$ are the solutions of 1.1.

First, we claim that for every $x \in X, A x \in X$. To see this, for every $t \in I$ and $x \in X$, we have

$$
\begin{aligned}
A x(t) & =\int_{0}^{T} G(t, s)[f(s, x(s))+\beta x(s)] d s \\
& =\int_{0}^{T} G(t, s) f(s, x(s)) d s+\int_{0}^{T} G(t, s) \beta x(s) d s \\
& >\int_{0}^{T} G(t, s) f(s, x(s)) d s+\beta \int_{0}^{T} G(t, s) d s \\
& \left.\left.=\int_{0}^{T} G(t, s) f(s, x(s)) d s+\beta \frac{1}{e^{\beta T}-1}\left(\frac{1}{\beta} e^{\beta(T+s-t)}\right]_{0}^{t}+\frac{1}{\beta} e^{\beta(s-t)}\right]_{t}^{T}\right) \\
& =\int_{0}^{T} G(t, s) f(s, x(s)) d s+\beta \frac{1}{\beta} \\
& =\int_{0}^{T} G(t, s) f(s, x(s)) d s+1
\end{aligned}
$$

one can conclude that $A x(t)>1$ and we have $A x \in X$.
Now, we check that the hypotheses in Theorem 2.1 is satisfied. To this end, we prove the following statements:

1. $A$ is $\perp$-preserving,
2. $A$ is $\perp$-contraction,
3. $A$ is $\perp$-continuous
4. We recall that $A$ is $\perp$-preserving if for every $x, y \in X, x \perp y$ we have $A x \perp A y$. We have shown above that $A x(t)>1$ for all $t \in I$, which implies that $A x(t) A y(t) \geq A y(t)$ for all $t \in I$. So $A x \perp A y$ if $x \perp y$.
5. Let $x, y \in X$ and $x \perp y$, we have

$$
\begin{aligned}
|A x(t)-A y(t)| & =\left|\int_{0}^{T} G(t, s)[f(s, x(s))+\beta x(s)-f(s, y(s))-\beta y(s)] d s\right| \\
& \leq \int_{0}^{T}|G(t, s) \| \mu(x(t)-y(t))| d s \\
& \leq \mu|x(t)-y(t)| \int_{0}^{T} G(t, s) d s \\
& =\frac{\mu}{\beta}|x(t)-y(t)|
\end{aligned}
$$

So,

$$
\begin{equation*}
d(A x, A y)=\sup _{t \in I}|A x(t)-A y(t)| \leq \frac{\mu}{\beta} \sup _{t \in I}|x(t)-y(t)|=\frac{\mu}{\beta} d(x, y) \tag{3.2}
\end{equation*}
$$

The inequality 3.2 shows that $A$ is $\perp$-contraction with Lipschitz constant $\lambda=\frac{\mu}{\beta}<1$.
3. Let $\left\{x_{n}\right\}$ be an O-sequence in $X$ such that $\left\{x_{n}\right\}$ converges to some $x \in X$. Since $A$ is $\perp$-preserving, $\left\{A x_{n}\right\}$ is an O-sequence. For each $n \in \mathbb{N}$, by (2), we have

$$
\left|A x_{n}-A x\right| \leq \lambda\left|x_{n}-x\right|
$$

As $n$ tends to infinity, it follows that $A$ is $\perp$-continuous.
The mapping $A$ defined above satisfies the hypotheses of the Theorem 2.1. Thus, the existence and uniqueness of its fixed point $x^{*} \in X$ has been guaranteed by Theorem 2.1. As we noted above, $x^{*}$ is a unique solution to differential equation 1.1.

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Zeinab Eivazi Damirchi Darsi Olia
Department of Mathematics
Islamic Azad University
South Tehran Branch, Tehran, Iran
St_z_eivazi@azad.ac.ir

Madjid Eshaghi Gordji
Department of Mathematics
Semnan University
P.O. Box 35195-363

Semnan, Iran
meshaghi@semnan.ac.ir

Davood Ebrahimi Bagha
Department of Mathematics
Islamic Azad University
Central Tehran Branch, Tehran, Iran
E_bagha@yahoo.com

# ARENS REGULARITY OF PROJECTIVE TENSOR PRODUCT 

Mostfa Shams Kojanaghi and Kazem Haghnejad Azar

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Abstract. In this paper, we study some of approximate identity properties, and its application in the Arens regularity of tensor products of Banach algebras with some results in group algebras. W e consider under which sufficient and necessary conditions the Banach algebra $A \widehat{\otimes} B$ is Arens regular.
Keywords: Arens regularity; tensor product; Banach algebras; group algebras.

## 1. Introduction

Suppose that $A$ and $B$ are Banach algebras. Since 1988 the Arens regularity of $A \widehat{\otimes} B$ has received a great deal of attention by many researchers. Among them, Ülger in [19, 21] showed that $A \widehat{\otimes} B$ is not Arens regular, in general, even when $A$ and $B$ are Arens regular. He introduced a new concept of biregular mapping and showed that a bounded bilinear mapping $m: A \times B \rightarrow \mathbb{C}$ is biregular if and only if $A \widehat{\otimes} B$ is Arens regular, where $\mathbb{C}$ is the space of complex numbers. Let $X, Y$ and $Z$ be normed spaces and let $m: X \times Y \rightarrow Z$ be a bounded bilinear mapping. Arens in [1] offers two natural extensions $m^{* * *}$ and $m^{t * * * t}$ of $m$ from $X^{* *} \times Y^{* *}$ into $Z^{* *}$ that he called $m$ is Arens regular whenever $m^{* * *}=m^{t * * * t}$, for more information see $[9,10,14]$. Let $A$ be a Banach algebra, regarding $A$ as a Banach $A$-bimodule, the operation $\pi: A \times A \longrightarrow A$ extends to $\pi^{* * *}$ and $\pi^{t * * * t}$ defined on $A^{* *} \times A^{* *}$. These extensions are known, respectively, as the first (left) and the second (right) Arens products, and with each of them, the second dual space $A^{* *}$ becomes a Banach algebra. The regularity of a normed algebra $A$ is defined to be the regularity of its algebra multiplication when considered as a bilinear mapping. The first (left) and second (right) Arens products of $a^{\prime \prime}, b^{\prime \prime} \in A^{* *}$ shall be simply indicated by $a^{\prime \prime} b^{\prime \prime}$ and $a^{\prime \prime} o b^{\prime \prime}$, respectively. Let $B$ be a Banach $A$-bimodule, and let

$$
\pi_{\ell}: A \times B \longrightarrow B \quad \text { and } \quad \pi_{r}: B \times A \longrightarrow B
$$

be the right and left module actions of $A$ on $B$. By above notation, the transpose of $\pi_{r}$ denoted by $\pi_{r}^{t}: A \times B \rightarrow B$. Then

$$
\pi_{\ell}^{*}: B^{*} \times A \longrightarrow B^{*} \quad \text { and } \quad \pi_{r}^{t * t}: A \times B^{*} \longrightarrow B^{*}
$$

Thus $B^{*}$ is a left Banach $A$-module and a right Banach $A$-module with respect to the module actions $\pi_{r}^{t * t}$ and $\pi_{\ell}^{*}$, respectively. The the second dual $B^{* *}$ is a Banach $A^{* *}$-bimodule with the following module actions

$$
\pi_{\ell}^{* * *}: A^{* *} \times B^{* *} \longrightarrow B^{* *} \quad \text { and } \pi_{r}^{* * *}: B^{* *} \times A^{* *} \longrightarrow B^{* *}
$$

where $A^{* *}$ is considered as a Banach algebra with respect to the first Arens product. Similarly, $B^{* *}$ is a Banach $A^{* *}$-bimodule with the module actions

$$
\pi_{\ell}^{t * * * t}: A^{* *} \times B^{* *} \longrightarrow B^{* *} \quad \text { and } \pi_{r}^{t * * * t}: B^{* *} \times A^{* *} \longrightarrow B^{* *}
$$

where $A^{* *}$ is considered as a Banach algebra with respect to the second Arens product.

Let $B$ be a left Banach $A$-module and $e$ be a left unit element of $A$. Then $e$ is a left unit (resp. weakly left unit) for $B$, if $\pi_{\ell}(e, b)=b$ (resp. $\left\langle b^{\prime}, \pi_{\ell}(e, b)\right\rangle=\left\langle b^{\prime}, b\right\rangle$ for all $b^{\prime} \in B^{*}$ ) where $b \in B$. The definition of right unit (resp. weakly right unit) is similar. A Banach $A$-bimodule $B$ is called unital, if $B$ has the same left and right unit. In this way, $B$ is called a unitary Banach $A$-bimodule.

Suppose that $A$ is a Banach algebra and $B$ is a Banach $A$-bimodule. Since $B^{* *}$ is a Banach $A^{* *}$-bimodule, where $A^{* *}$ is equipped with the first Arens product, we define the topological center of the right module action of $A^{* *}$ on $B^{* *}$ as follows:

$$
\begin{aligned}
Z_{A^{* *}}^{\ell}\left(B^{* *}\right)=Z\left(\pi_{r}\right)= & \left\{b^{\prime \prime} \in B^{* *}: \text { the map } a^{\prime \prime} \rightarrow \pi_{r}^{* * *}\left(b^{\prime \prime}, a^{\prime \prime}\right): A^{* *} \rightarrow B^{* *}\right. \\
& \text { is weak*-weak } \left.{ }^{*} \text { continuous }\right\} .
\end{aligned}
$$

In this way, we write $Z_{B^{* *}}^{\ell}\left(A^{* *}\right)=Z\left(\pi_{\ell}\right), Z_{A^{* *}}^{r}\left(B^{* *}\right)=Z\left(\pi_{\ell}^{t}\right)$ and $Z_{B^{* *}}^{r}\left(A^{* *}\right)=$ $Z\left(\pi_{r}^{t}\right)$, where $\pi_{\ell}: A \times B \rightarrow B$ and $\pi_{r}: B \times A \rightarrow B$ are the left and right module actions of $A$ on $B$, for more information related to the Arens regularity of module actions on Banach algebras, see $[2,4,9,10]$. If we set $B=A$, we write $Z_{A^{* *}}^{\ell}\left(A^{* *}\right)=Z_{1}\left(A^{* *}\right)=Z_{1}^{\ell}\left(A^{* *}\right)$ and $Z_{A^{* *}}^{r}\left(A^{* *}\right)=Z_{2}\left(A^{* *}\right)=Z_{2}^{r}\left(A^{* *}\right)$, for more information see [12]. Let $A$ be a Banach algebra, $A^{*}$ and $A^{* *}$ be the first and second dual of $A$, respectively. For $a \in A$ and $a^{\prime} \in A^{*}$, by $a^{\prime} a$ and $a a^{\prime}$, we mean the functionals in $A^{*}$ defined by $\left\langle a^{\prime} a, b\right\rangle=\left\langle a^{\prime}, a b\right\rangle=a^{\prime}(a b)$ and $\left\langle a a^{\prime}, b\right\rangle=\left\langle a^{\prime}, b a\right\rangle=a^{\prime}(b a)$ for all $b \in A$, respectively. A Banach algebra $A$ is embedded in its second dual via the identification $\left\langle a, a^{\prime}\right\rangle-\left\langle a^{\prime}, a\right\rangle$ for every $a \in A$ and $a^{\prime} \in A^{*}$.

## 2. Main Results

Consider the tensor product, $X \otimes Y$, of the vector space $X$ and $Y$ which can be constructed as a space of linear functional on $B(X \times Y)$. By $X \widehat{\otimes} Y$ we shall denote the projective tensor products of $X$ and $Y$, where $X \widehat{\otimes} Y$ is the completion of $X \otimes Y$ for the norm

$$
\|u\|=\inf \sum_{i=1}^{n}\left\|x_{i}\right\|\left\|y_{i}\right\|
$$

where the infimum is taken over all the representations of $u$ as a finite sum of the form $u=\sum_{i=1}^{n} x_{i} \otimes y_{i}[5]$.

The natural multiplication of $A \widehat{\otimes} B$ is the linear extension of the following multiplication on decomposable tensors $(a \otimes b)(\tilde{a} \otimes \tilde{b})=a \tilde{a} \otimes b \tilde{b}$. For more details, see [16].

A functional $a^{\prime}$ in $A^{*}$ is said to be wap (weakly almost periodic) on $A$ if the mapping $a \rightarrow a^{\prime} a$ from $A$ into $A^{*}$ is weakly compact. Pym in [15] showed that this definition is equivalent with the following condition:

$$
\lim _{i} \lim _{j}\left\langle a^{\prime}, a_{i} b_{j}\right\rangle=\lim _{j} \lim _{i}\left\langle a^{\prime}, a_{i} b_{j}\right\rangle,
$$

whenever both iterated limits exist, for any two net $\left(a_{i}\right)_{i}$ and $\left(b_{j}\right)_{j}$ in $\{a \in A$ : $\|a\| \leq 1\}$. The collection of all weakly almost periodic functionals on $A$ is denoted by $\operatorname{wap}(A)$. Also, $a^{\prime} \in \operatorname{wap}(A)$ if and only if $\left\langle a^{\prime \prime} b^{\prime \prime}, a^{\prime}\right\rangle=\left\langle a^{\prime \prime} o b^{\prime \prime}, a^{\prime}\right\rangle$ for every $a^{\prime \prime}, b^{\prime \prime} \in A^{* *}$. Thus, it is clear that $A$ is Arens regular if and only if $\operatorname{wap}(A)=A^{*}[9$, Theorem 2.6.17]. In the sequel, to show that the projective tensor products $A \widehat{\otimes} B$ is Arens regular, it is sufficient that we show that $\operatorname{wap}(A \widehat{\otimes} B)=(A \widehat{\otimes} B)^{*}$. In all of this section, we regard $A^{*} \widehat{\otimes} B^{*}$ as a subset of $(A \widehat{\otimes} B)^{*}$ and by $A_{1}$ and $B_{1}$ we mean all elements of $a \in A$ and $b \in B$ such that $\|a\| \leq 1$ and $\|b\| \leq 1$.

Theorem 2.1. Suppose that $A$ and $B$ are Banach algebras and for every sequence $\left(x_{i}\right)_{i},\left(y_{j}\right)_{j} \subseteq A_{1},\left(z_{i}\right)_{i},\left(w_{j}\right)_{j} \subseteq B_{1}$ and $f \in B(A \times B)$, we have

$$
\lim _{j} \lim _{i} f\left(x_{i} z_{i}, y_{j} w_{j}\right)=\lim _{i} \lim _{j} f\left(x_{i} z_{i}, y_{j} w_{j}\right) .
$$

Then $A \widehat{\otimes} B$ is Arens regular.
Proof. Assume that $f \in B(A \times B)$. Since $B(A \times B)=(A \widehat{\otimes} B)^{*}$, it is enough to show that $f \in \operatorname{wap}(A \widehat{\otimes} B)$. Let $\left(x_{i}\right)_{i},\left(y_{j}\right)_{j} \subseteq A_{1}$ and $\left(z_{i}\right)_{i},\left(w_{j}\right)_{j} \subseteq B_{1}$, then we have the following equality

$$
\begin{aligned}
\lim _{j} \lim _{i}\left\langle f,\left(x_{i} \otimes y_{j}\right)\left(z_{i} \otimes w_{j}\right)\right\rangle & =\lim _{j} \lim _{i}\left\langle f, x_{i} z_{i} \otimes y_{j} w_{j}\right\rangle \\
& =\lim _{j} \lim _{i} f\left(x_{i} z_{i}, y_{j} w_{j}\right) \\
& =\lim _{i} \lim _{j} f\left(x_{i} z_{i}, y_{j} w_{j}\right) \\
& =\lim _{i} \lim _{j}\left\langle f,\left(x_{i} \otimes y_{j}\right)\left(z_{i} \otimes w_{j}\right)\right\rangle
\end{aligned}
$$

for every $f \in(A \widehat{\otimes} B)^{*}$. This means that $f \in \operatorname{wap}(A \widehat{\otimes} B)$, and proof is complete.

Definition 2.1. Let $A$ be a Banach algebra and let $B$ be a Banach $A$-bimodule and let $\pi: A \widehat{\otimes} B \longrightarrow B$ such that $\pi(a \otimes b)=a b$ for every $a \in A, b \in B$. We say that $B$ is non-trivial on $A$, if $\pi$ is surjective and has a bounded right inverse.

Remark 2.1. In the above definition, if $A$ is unital, then $\pi$ will be surjective. Now, suppose $\pi$ has a continuous right inverse $\rho, e_{A}$ and $e_{B}$ are units of $A$ and $B$, respectively. Let $\varphi \in(A \widehat{\otimes} B)^{*}$, then $\varphi \circ \rho$ belongs to $B^{*}$. Hence, there is a $\phi \in B^{*}$ such that $\varphi \circ \rho=\phi$. In other word, in the following diagram, we have $\varphi \circ \rho=\phi \circ \operatorname{id}_{B}$.


As well as, $\phi \circ \pi$ is in $(A \widehat{\otimes} B)^{*}$. Thus, there is a $\psi \in(A \widehat{\otimes} B)^{*}$ such that $\phi \circ \pi=\psi$. Then $\phi=\psi \circ \rho$. For given $a \otimes b \in A \widehat{\otimes} B$ we have

$$
\begin{align*}
\phi \circ \pi(a \otimes b) & =\phi(a b)=\varphi \circ \rho(a b)=\varphi \circ \rho\left(e_{A} a e_{B} b\right) \\
& =\varphi \circ \rho\left(\left(e_{A} a\right)\left(e_{B} b\right)\right)=\psi \circ \rho\left(\left(e_{A} a\right)\left(e_{B} b\right)\right) \\
& =\psi(a \otimes b) . \tag{2.1}
\end{align*}
$$

Then, by (2.1), for every $\varphi \in(A \widehat{\otimes} B)^{*}$ there is a $\psi \in(A \widehat{\otimes} B)^{*}$ such that $\varphi \circ \rho=\psi \circ \rho$ and $\psi(a \otimes b)=\psi \circ \rho(a b)$, for every $a \in A$ and $b \in B$. Since $A$ is unital, every element $c$ of $B$ can be written as $c=a b$ where $a \in A$ and $b \in B$. We can define $\rho: B \longrightarrow A \widehat{\otimes} B$ by $\rho(b)=e_{A} \otimes b$ and $\rho(a b)=\rho\left(\left(e_{A} a\right) b\right)=a \otimes b$, for every $a \in A$ and $b \in B$. By this definition $\rho$ is injective and it is a unique way to define of $\rho$. By this definition the above diagram commutes and we have $\phi \circ \pi(a \otimes b)=\varphi(a \otimes b)$, for every $a \in A$ and $b \in B$.

A wide class of Banach algebras which satisfy in the Definition 2.1, are projective and biprojective Banach algebras. A Banach algebra $A$-bimodule $B$ is called projective if $\pi: A^{\sharp} \widehat{\otimes} B \longrightarrow B$ has bounded right inverse in ${ }_{A} B\left(B, A^{\sharp} \widehat{\otimes} B\right)$ and the Banach algebra $A$ is called biprojective if $\pi: A \widehat{\otimes} A \longrightarrow A$ has bounded right inverse in ${ }_{A} B(A, A \widehat{\otimes} A$ ) (for more details see [18]).

Theorem 2.2. Let $A$ and $B$ be Banach algebras and $B$ is unital. Suppose $B$ is a Banach A-bimodule. Then

1. if $A \widehat{\otimes} B$ is Arens regular, then $A$ is Arens regular.
2. if $B$ is non-trivial on $A$ and $B$ be a unitary Banach $A$-bimodule. Then $A$ and $B$ are Arens regular if and only if $A \widehat{\otimes} B$ is Arens regular.

Proof. 1. Assume that $A \widehat{\otimes} B$ is Arens regular and $u \in B$ is the unit element of $B$. We show that $\operatorname{wap}(A)=A^{*}$. Get $\left(a_{i}\right)_{i} \subseteq A,\left(c_{j}\right)_{j} \subseteq A$ and $a^{\prime} \in A^{*}$. Define $\phi=a^{\prime} \otimes b^{\prime}$ where $b^{\prime} \in B^{*}$ and $b^{\prime}(u)=1$. Since $A^{*} \otimes B^{*} \subseteq(A \widehat{\otimes} B)^{*}$ and $A \widehat{\otimes} B$ is Arens regular, we have $a^{\prime} \otimes b^{\prime} \in \operatorname{wap}(A \widehat{\otimes} B)$. Hence it follows that

$$
\begin{aligned}
\lim _{i} \lim _{j}\left\langle a^{\prime}, a_{i} c_{j}\right\rangle & =\lim _{i} \lim _{j}\left\langle a^{\prime} \otimes b^{\prime}, a_{i} c_{j} \otimes u\right\rangle \\
& =\lim _{i} \lim _{j}\left\langle a^{\prime} \otimes b^{\prime},\left(a_{i} \otimes u\right)\left(c_{j} \otimes u\right)\right\rangle \\
& =\lim _{j} \lim _{i}\left\langle a^{\prime} \otimes b^{\prime},\left(a_{i} \otimes u\right)\left(c_{j} \otimes u\right)\right\rangle \\
& =\lim _{j} \lim _{i}\left\langle a^{\prime}, a_{i} c_{j}\right\rangle .
\end{aligned}
$$

This means that $a^{\prime} \in \operatorname{wap}(A)$, and so $A$ is Arens regular.
2. Let $u$ be a unit element of $B$ and let $B$ be Arens regular. Then $\operatorname{wap}(B)=B^{*}$. Suppose that $\left(a_{i}\right)_{i} \subseteq A_{1}$ and $\left(b_{j}\right)_{j} \subseteq B_{1}$ whenever both iterated limits exist. Then $\left(a_{i} u\right)_{i} \subseteq B_{1}$, and so for every $b^{\prime} \in B^{*}$, we have the following equality

$$
\lim _{i} \lim _{j}\left\langle b^{\prime},\left(a_{i} u\right) b_{j}\right\rangle=\lim _{j} \lim _{i}\left\langle b^{\prime},\left(a_{i} u\right) b_{j}\right\rangle .
$$

Now; let $\varphi \in(A \widehat{\otimes} B)^{*}$. Then by Remark 2.1, $\pi: A \widehat{\otimes} B \longrightarrow B$ has a continuous right inverse $\rho$ such that $\varphi \circ \rho$ belongs to $B^{*}$ and there is a $\phi \in B^{*}$ such that $\varphi \circ \rho=\phi$, and $\phi \circ \pi(a \otimes b)=\varphi(a \otimes b)$, for every $a \otimes b \in A \widehat{\otimes} B$. Now we have

$$
\begin{aligned}
\lim _{i} \lim _{j}\left\langle\varphi, a_{i} \otimes b_{j}\right\rangle & =\lim _{i} \lim _{j}\left\langle\phi \circ \pi, a_{i} \otimes b_{j}\right\rangle=\lim _{i} \lim _{j}\left\langle\phi, \pi\left(a_{i} \otimes b_{j}\right)\right\rangle \\
& =\lim _{i} \lim _{j}\left\langle\phi, a_{i} b_{j}\right\rangle=\lim _{i} \lim _{j}\left\langle\phi, a_{i}\left(u b_{j}\right)\right\rangle \\
& =\lim _{j} \lim _{i}\left\langle\phi,\left(a_{i} u\right) b_{j}\right\rangle=\lim _{j} \lim _{i}\left\langle\phi, \pi\left(a_{i} \otimes b_{j}\right)\right\rangle \\
& =\lim _{j} \lim _{i}\left\langle\varphi, a_{i} \otimes b_{j}\right\rangle .
\end{aligned}
$$

It follows that $\varphi \in \operatorname{wap}(A \widehat{\otimes} B)$, and so $A \widehat{\otimes} B$ is Arens regular. The converse by using the part (1) holds.

Corollary 2.1. Suppose that $A$ and $B$ are unital Banach algebras and $B$ is a unitary Banach A-bimodule. Assume that $B$ is non-trivial on $A$. If $A$ and $B$ are Arens regular, then every bilinear form $m: A \times B \rightarrow \mathbb{C}$ is weakly compact.

Proof. Apply Theorem 2.2 and Theorem 3.4 of [19].
Let $A$ and $B$ be Banach algebras. A bilinear form $m: A \times B \rightarrow \mathbb{C}$ is said to be biregular, if for any two pairs of sequence $\left(a_{i}\right)_{i},\left(\tilde{a}_{j}\right)_{j}$ in $A_{1}$ and $\left(b_{i}\right)_{i},\left(\tilde{b}_{j}\right)_{j}$ in $B_{1}$, we have

$$
\lim _{i} \lim _{j} m\left(a_{i} \tilde{a}_{j}, b_{i} \tilde{b}_{j}\right)=\lim _{j} \lim _{i} m\left(a_{i} \tilde{a}_{j}, b_{i} \tilde{b}_{j}\right)
$$

provided that these limits exist. There are some examples of biregular non regular bilinear form for more information see [19].

Corollary 2.2. Suppose that $A$ and $B$ are Banach algebras. Then we have the following assertions:

1. By the conditions of Theorem 2.1, every bilinear form $m: A \times B \rightarrow \mathbb{C}$ is biregular.
2. By the conditions of Theorem 2.2 (2), every bilinear form $m: A \times B \rightarrow \mathbb{C}$ is biregular.

In the following, we give a simple proof of Theorem 3.4 of [19].
Theorem 2.3. [19, Theorem 3.4] Let $A$ and $B$ be Banach algebras and $u: A \rightarrow B^{*}$ be a continuous linear operator. Then the bilinear form $m: A \times B \rightarrow \mathbb{C}$ defined by $m(a, b)=\langle u(a), b\rangle$ is biregular.

Proof. Let $\left(a_{i}\right)_{i},\left(\tilde{a}_{j}\right)_{j}$ in $A_{1}$ and $\left(b_{i}\right)_{i},\left(\tilde{b}_{j}\right)_{j}$ in $B_{1}$ such that the following iterated limits exist:

$$
\lim _{i} \lim _{j} m\left(a_{i} \tilde{a}_{j}, b_{i} \tilde{b}_{j}\right) \text { and } \lim _{j} \lim _{i} m\left(a_{i} \tilde{a}_{j}, b_{i} \tilde{b}_{j}\right) \text {. }
$$

There are $\left(a_{\alpha}\right)_{\alpha},\left(\tilde{a}_{\beta}\right)_{\beta}$ in $A$ and $\left(b_{\alpha}\right)_{\alpha},\left(\tilde{b}_{\beta}\right)_{\beta}$ in $B$ such that $a_{\alpha} \xrightarrow{w^{*}} a^{\prime \prime}$ and $\tilde{a}_{\beta} \xrightarrow{w^{*}} \tilde{a}^{\prime \prime}$ in $A^{* *}$ and we have $b_{\alpha} \xrightarrow{w^{*}} b^{\prime \prime}$ and $\tilde{b}_{\beta} \xrightarrow{w^{*}} \tilde{b}^{\prime \prime}$ in $B^{* *}$. Since $A$ and $B$ are Arens regular, we have

$$
\lim _{\alpha} \lim _{\beta} a_{\alpha} \tilde{a}_{\beta}=\lim _{\beta} \lim _{\alpha} a_{\alpha} \tilde{a}_{\beta}=a^{\prime \prime} \tilde{a}^{\prime \prime}
$$

and

$$
\lim _{\alpha} \lim _{\beta} b_{\alpha} \tilde{b}_{\beta}=\lim _{\beta} \lim _{\alpha} b_{\alpha} \tilde{b}_{\beta}=b^{\prime \prime} \tilde{b}^{\prime \prime}
$$

Then, since $u$ is continuous, we have

$$
\begin{aligned}
\lim _{\alpha} \lim _{\beta} m\left(a_{\alpha} \tilde{a}_{\beta}, b_{\alpha} \tilde{b}_{\beta}\right) & =\lim _{\alpha} \lim _{\beta}\left\langle u\left(a_{\alpha} \tilde{a}_{\beta}\right), b_{\alpha} \tilde{b}_{\beta}\right\rangle \\
& =\left\langle u^{\prime \prime}\left(a^{\prime \prime} \tilde{a}^{\prime \prime}\right), b^{\prime \prime} \tilde{b}^{\prime \prime}\right\rangle .
\end{aligned}
$$

Similarly, we have

$$
\lim _{\beta} \lim _{\alpha} m\left(a_{\alpha} \tilde{a}_{\beta}, b_{\alpha} \tilde{b}_{\beta}\right)=\left\langle u^{\prime \prime}\left(a^{\prime \prime} \tilde{a}^{\prime \prime}\right), b^{\prime \prime} \tilde{b}^{\prime \prime}\right\rangle
$$

Consequently, we have

$$
\lim _{i} \lim _{j} m\left(a_{i} \tilde{a}_{j}, b_{i} \tilde{b}_{j}\right)=\lim _{j} \lim _{i} m\left(a_{i} \tilde{a}_{j}, b_{i} \tilde{b}_{j}\right) .
$$

It follows that $m$ is biregular.

Example 2.1. [19] Let $A$ be a Banach algebra and $1<p<\infty$. Then

1. $\ell^{p} \widehat{\otimes} A$ is Arens regular if and only if $A$ is Arens regular.
2. Let $G$ be a locally compact group. Then, $L^{p}(G) \widehat{\otimes} A$ is Arens regular if and only if $A$ is Arens regular.

Proof. For prove, we apply Theorem 3.4 of [19] and Theorem 2.7.
We finish this section with the following problems:
Problem 2.1. Let $G$ be a locally compact group. What can say for the following sets?

$$
Z_{L^{1}(G)^{* *}}^{\ell}\left(\left(L^{1}(G) \widehat{\otimes} L^{1}(G)\right)^{* *}\right)=? \quad Z_{L^{1}(G)^{* *}}^{\ell}\left(L^{1}(G)^{* *} \widehat{\otimes} L^{1}(G)^{* *}\right)=?
$$

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Mostfa Shams Kojanaghi
Department of Mathematics
Islamic Azad University Tehran
Tehran, Iran
mstafa.shams99@yahoo.com

Kazem Haghnejad Azar
Faculty of Science
Department of Mathematics
University of Mohaghegh Ardabili, Ardabil, Iran
Ardabil, Iran
haghnejad@uma.ac.ir

# STRONG CONVERGENCE THEOREM FOR UNIFORMLY L-LIPSCHITZIAN MAPPING OF GREGUS TYPE IN BANACH SPACES 

Olilima O. Joshua, Mogbademu A. Adesanmi and Adeniran T. Adefemi

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Abstract. In this paper, we introduced a new mapping called uniformly L-Lipschitzian mapping of Gregus type, and used the Mann iterative scheme to approximate the fixed point. A Strong convergence result for the sequence generated by the scheme is shown in real Banach space. Our result generalized and unify many recent results in this area of research. In addition, using Java (jdk 1.8.0_101), we give a numerical example to support our claim.
Key words: Mann iterative scheme; uniformly L-Lipschitzian mapping; normalized duality mapping.

## 1. Introduction

Let $E$ and $E^{*}$ be a real Banach space and its dual space respectively. The normalized duality mapping $J: E \rightarrow 2^{E^{*}}$ is defined by

$$
J(x)=\left\{f \in E^{*}:\langle x, f\rangle=\|x\|^{2}=\|f\|^{2},\|x\|=\|f\|\right\}
$$

where $\langle\cdot, \cdot\rangle$ denotes the duality pairing.
Definition 1.1. [Ofoedu E.U [13]] Let $K$ be a nonempty closed convex subset of a real Banach space $E$. The mapping $T: K \rightarrow E$ is said to be
i) nonexpansive if for all $x, y \in K$

$$
\begin{equation*}
\|T x-T y\| \leqslant\|x-y\| . \tag{1.1}
\end{equation*}
$$

ii) uniformly L-Lipschitzian if there exists $L>0$ such that, for any $x, y \in K$

$$
\begin{equation*}
\left\|T^{n} x-T^{n} y\right\| \leqslant L\|x-y\|, \quad \forall n \geqslant 1 . \tag{1.2}
\end{equation*}
$$

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iii) asymptotically nonexpansive if there exists $k_{n} \subset[1, \infty)$ with $k_{n} \rightarrow 1$ as $n \rightarrow \infty$ such that for any given $x, y \in K$,

$$
\begin{equation*}
\left\|T^{n} x-T^{n} y\right\| \leqslant k_{n}\|x-y\|, \quad \forall n>1 \tag{1.3}
\end{equation*}
$$

iv) asymptotically pseudocontractive if there exists a sequence $k_{n} \subset[0, \infty)$ with $k_{n} \rightarrow 1$ as $n \rightarrow \infty$. and there exists $j(x-y) \in J(x-y)$ such that

$$
\begin{equation*}
\left\langle T^{n} x-T^{n} y, j(x-y)\right\rangle \leqslant k_{n}\|x-y\|^{2}, \quad n \geqslant 1 \tag{1.4}
\end{equation*}
$$

We can easily see from equations (1.2), (1.3), (1.4) that the class of asymptotically non-expansive mappings is a generalization of the class of uniformly L-Lipschitzian mapping. And that every asymptotically nonexpansive mappings are asymptotically pseudocontractive, the reason is shown below,

$$
\begin{equation*}
\left\langle T^{n} x-T^{n} y, j(x-y)\right\rangle \leqslant\left\|T^{n} x-T^{n} y\right\|\|x-y\| \leqslant k_{n}\|x-y\|^{2}, \quad \forall n \geqslant 1 \tag{1.5}
\end{equation*}
$$

But the converse is not always true. The example to show that the converse is not true was constructed by Rhoades [15]. The asymptotically nonexpansive mappings and the asymptotically pseudocontractive mappings were introduced by Goebel and Kirk [4] and Schu [16] respectively.

In 1980, Gregus [5] introduced what is now known as the Gregus fixed point theorem. He proved the following theorem.

Theorem 1.1. Gregus [5] Let $K$ be a closed convex subset of a Banach space $E$ and $T: K \rightarrow K$ a mapping that satisfies $\|T x-T y\| \leqslant a\|x-y\|+b\|x-T x\|+c\|y-T y\|$ for all $x, y \in K$ where $0<a<1, b, c \geqslant 0$ and $a+b+c=1$. Then $T$ has a unique fixed point.

The class of mapping introduced by Gregus [5] is a generalization of nonexpansive mapping which is a very important mapping in fixed point theorem and applications, because if $a=1, b=c=0$ then we have the mapping in (1.1), and if $a=0, b=c=\frac{1}{2}$ we have the Kannan mappings introduced by Kannan in [6]. This class of mappings have been extended by many authors in various ways and under different conditions on $T$. For results on these, see $[8,11,12,14]$ and the references therein.

The trend for uniformly L-Lipschitzian mapping and asymptotically pseudocontractive mapping is given below for better understanding the concept we intend to introduce.

In 1991 Schu [16], proved the following result using the modified Mann iterative scheme

Theorem 1.2. Schu [16] Let $H$ be a Hilbert space, $K$ be a nonempty bounded closed convex subset of $H$ and $T: K \rightarrow K$ be a completely continuous, uniformly $L$ Lipschitzian and asymptotically pseudo-contractive mapping with a sequence $\left\{k_{n}\right\}$ satisfying the following conditions:
(i) $k_{n} \rightarrow 1$ as $n \rightarrow 1$;
(ii) $\sum_{n=1}^{\infty} q_{n}^{2}-1<\infty$, where $q_{n}=2 k_{n}-1$.

Suppose further that $\left\{\alpha_{n}\right\}$ and $\left\{\beta_{n}\right\}$ are two sequences in $[0,1]$ such that $\epsilon<\alpha_{n}<$ $b, \quad \forall n \geqslant 1$ where $\epsilon>0$ and $b \in\left(0, L^{-2}\left[\left(1+L^{2}\right)^{1 / 2}-1\right]\right)$ are some positive numbers. For any $x_{1} \in K$, let $\left\{x_{n}\right\}$ be the iterative sequence defined by

$$
\begin{equation*}
x_{n+1}=\left(1-\alpha_{n}\right) x_{n}+\alpha_{n} T^{n} x_{n}, \quad \forall n \geqslant 1 . \tag{1.6}
\end{equation*}
$$

Then $\left\{x_{n}\right\}$ converges strongly to a fixed point of $T$ in $K$.
In 2000, Chang [1] extended theorem 1.2 from Hilbert space to uniformly smooth Banach space, by proving the following theorem:

Theorem 1.3. Chang [1] Let $E$ be a real uniformly smooth Banach space, $K$ be a nonempty bounded closed convex subset of $E, T: K \rightarrow K$ be an asymptotically pseudocontractive mapping with sequence $\left\{k_{n}\right\} \subset[1,+\infty), \lim _{n \rightarrow \infty} k_{n}=1$, and $F(T)=$ $\{x \in K: T x=x\} \neq \varnothing$. Let $\alpha_{n} \subset[0,1]$ satisfying the following conditions:
(i) $\alpha_{n} \rightarrow 0$ as $n \rightarrow \infty$;
(ii) $\sum_{n=1}^{\infty} \alpha_{n}=\infty$.

For any $x_{0} \in K$, let $\left\{x_{n}\right\}$ be the iterative sequence defined by

$$
x_{n+1}=\left(1-\alpha_{n}\right) x_{n}+\alpha_{n} T^{n} x_{n}, \quad \forall n \geqslant 0,
$$

If there exists a strict increasing function $\Phi:[0, \infty) \rightarrow[0, \infty)$ with $\Phi(0)=0$ such that

$$
\left\langle T^{n} x-p, j(x-p)\right\rangle \leqslant k_{n}\|x-p\|^{2}-\Phi(\|x-p\|)
$$

for all $x \in K$ and $n \geqslant 0$, where $p \in F(T)$, then $x_{n} \rightarrow p$ as $n \rightarrow \infty$.
Recently Ofoedu [13], extended theorem 1.3 from uniformly smooth Banach space to real Banach space and he also dispensed with the boundedness condition imposed by earlier researchers, by stating and proving the following theorem:

Theorem 1.4. Ofoedu [13] Let $E$ be a real Banach space, $K$ be a nonempty closed and convex subset of $E, T: K \rightarrow K$ be a uniformly L-Lipschitzian asymptotically pseudocontractive mapping with sequence $\left\{k_{n}\right\}_{n \geqslant 0} \subset[1,+\infty), \lim _{n \rightarrow \infty} k_{n}=1$, and let $p \in F(T)=\{x \in K: T x=x\} \neq \varnothing$. Let $\alpha_{n} \subset[0,1]$ satisfying the following conditions:
(i) $\sum_{n=1}^{\infty} \alpha_{n}=\infty$;
(ii) $\sum_{n=1}^{\infty} \alpha_{n}^{2}<\infty$;
(iii) $\sum_{n=1}^{\infty} \alpha_{n}\left(k_{n}-1\right)<\infty$

For any $x_{0} \in K$, let $\left\{x_{n}\right\}$ be the iterative sequence defined by

$$
x_{n+1}=\left(1-\alpha_{n}\right) x_{n}+\alpha_{n} T^{n} x_{n}, \quad \forall n \geqslant 0 .
$$

If there exists a strict increasing function $\Phi:[0, \infty) \rightarrow[0, \infty)$ with $\Phi(0)=0$ such that

$$
\left\langle T^{n} x-p, j(x-p)\right\rangle \leqslant k_{n}\|x-p\|^{2}-\Phi(\|x-p\|)
$$

for all $x \in K$ and $n \geqslant 0$, where $p \in F(T)$, then $x_{n} \rightarrow p$ as $n \rightarrow \infty$.
Inspired by the above results, we introduce the following concept which generalizes the class of uniformly L-Lipschitzian mappings. The mapping is defined as follows:

Definition 1.2. Let $K$ be a nonempty closed subset of a real Banach space $E$. The mapping $T: K \rightarrow E$ is said to be uniformly L-Lipschitzian mapping of Gregus type if there exists $L>0$, and the sequences $a_{n}, b_{n} \in[0, \infty)$, with $a_{n}, b_{n} \rightarrow 0$ as $n \rightarrow \infty$, such that for any $x, y \in K$,

$$
\begin{equation*}
\left\|T^{n} x-T^{n} y\right\| \leqslant L\|x-y\|+a_{n}\left\|x-T^{n} x\right\|+b_{n}\left\|y-T^{n} y\right\|, \quad \forall n \geqslant 1 \tag{1.7}
\end{equation*}
$$

If we set $a_{n}=b_{n}=0 \quad \forall n \in \mathbb{N}$, equation (1.7) is reduced to (1.2). Clearly, every uniformly L -Lipschitzian mapping is uniformly L -Lipschitaian mapping of Gregus type, but the converse is not generally true. It suffices to construct an example of a map that is uniformly L-Lipschitzian of Gregus type but not uniformly L-Lipschitzian.

Example 1.1. Let $E=\mathbb{R}$ be the set of real numbers with the usual norm, and let $K=[0, \infty)$. Consider the mapping $T: K \rightarrow K$ defined by

$$
T x=\frac{x^{3}}{4(1+x)}, \quad \forall x \in K
$$

It is easy to see that $T$ is a monotone increasing function satisfying (1.7), but $T$ does not satisfy inequality (1.2). In fact,

$$
\begin{aligned}
\left|T^{n} x-T^{n} y\right| & \leqslant|T x-T y|=\left|\frac{x^{3}}{4+4 x}-\frac{y^{3}}{4+4 y}\right|=\frac{1}{4}\left|\frac{x^{3}}{1+x}-\frac{y^{3}}{1+y}\right| \\
& =\frac{1}{4}\left|\frac{x-y+x^{3}(1+y)-x+y-y^{3}(1+x)}{(1+x)(1+y)}\right| \\
& \leqslant \frac{1}{4}\left|\frac{x-y}{(1+x)(1+y)}\right|+\left|\frac{x^{3}(1+y)-x}{4(1+x)(1+y)}\right|+\left|\frac{y-y^{3}(1+x)}{4(1+x)(1+y)}\right| \\
& \leqslant \frac{1}{4}|x-y|+\left|x-\frac{x^{3}}{4(1+x)}\right|+\left|y-\frac{y^{3}}{4(1+y)}\right| \\
& =\frac{1}{4}|x-y|+|x-T x|+|y-T y| .
\end{aligned}
$$

Therefore,

$$
\begin{equation*}
\left|T^{n} x-T^{n} y\right| \leqslant \frac{1}{4}|x-y|+|x-T x|+|y-T y| . \tag{1.8}
\end{equation*}
$$

Hence, $T$ is uniformly L-Lipschitzian mapping of Gregus type where the sequences $a_{n}=$ $b_{n}=1, \forall n \in \mathbb{N}$ and $L=\frac{1}{4}$. But observe that,

$$
\frac{x^{3}}{4(1+x)}>x \quad \forall x>\frac{4+\sqrt{32}}{2},
$$

hence we have that,

$$
\left|T^{n} x-T^{n} y\right| \leqslant|T x-T y|=\left|\frac{x^{3}}{4+4 x}-\frac{y^{3}}{4+4 y}\right|>|x-y| .
$$

Thus, $T$ is not a uniformly L-Lipschitzian mapping. We can now say that the class of uniformly L- Lipschitzian mappings of Gregus type properly includes the class of uniformly L-Lipschitzian mappings. Hence, it is more important to study this class of mappings in fixed point theory and applications.

In particular, If we let $x$ to be any point in $K$ and $y \in F(T)$ then, from (1.8) we have,

$$
\left|T^{n} x-T^{n} y\right| \leqslant \frac{1}{2}|x-0|+\frac{1}{4}|x-T x|+\frac{1}{4}|0-T 0| .
$$

but

$$
\left|T^{n} x-T^{n} y\right|=\left|T^{n} x-0\right| \leqslant|T x-0|=\left|\frac{x^{3}}{4+4 x}-0\right|>|x-0|,
$$

for $x>\frac{4+\sqrt{32}}{2}, y=0 \in F(T)$.
It is our aim in this paper to consider the iterative scheme in (1.6) and prove a strong convergence theorem for the newly introduced uniformly L-Lipschitzian mappings of Gregus type to a unique fixed point in real Banach spaces.

## 2. Preliminaries

We shall need the following Proposition and lemmas in the main theorem.
Proposition 2.1. Let $K$ be a nonempty closed convex subset of a Banach space and $T: K \rightarrow K$ be a Uniformly L-Lipschitzian mapping of Gregus Type with $a_{n} \in\left(0, \frac{1}{2}\right)$ such that $a_{n} \rightarrow 0$ as $n \rightarrow \infty$. Then for any $x_{0} \in K$, let $\left\{x_{n}\right\}$ be an iterative sequence defined by

$$
x_{n+1}=\left(1-\alpha_{n}\right) x_{n}+\alpha_{n} T^{n} x_{n}
$$

for all $n \geqslant 0$, there exists an $M>0$ such that for $p \in F(T)$, the following hold:
i.) $\left\|T^{n} x_{n}-x_{n}\right\| \leqslant M\left\|x_{n}-p\right\|$
ii.) $\left\|T^{n} x_{n+1}-x_{n+1}\right\| \leqslant M\left\|x_{n+1}-p\right\|$

Proof. Since $T$ is uniformly L-Lipschitzian mapping of Gregus Type, we have,

$$
\begin{align*}
\left\|x_{n}-T^{n} x_{n}\right\| & \leqslant\left\|x_{n}-p\right\|+\left\|T^{n} x_{n}-p\right\|=\left\|x_{n}-p\right\|+\left\|T^{n} x_{n}-T^{n} p\right\| \\
& \leqslant\left\|x_{n}-p\right\|+L\left\|x_{n}-p\right\|+a_{n}\left\|x_{n}-T^{n} x_{n}\right\|+b_{n}\left\|p-T^{n} p\right\| \\
& =(1+L)\left\|x_{n}-p\right\|+a_{n}\left\|x_{n}-T^{n} x_{n}\right\| \\
& \leqslant \frac{(1+L)}{1-a_{n}}\left\|x_{n}-p\right\| . \tag{2.1}
\end{align*}
$$

Since, $a_{n} \in[0,1 / 2)$ we have that, $-a_{n}>-\frac{1}{2}$. Hence, $1-a_{n}>1-\frac{1}{2}$, this implies that

$$
\frac{1}{1-a_{n}}<2
$$

Therefore,

$$
\frac{(1+L)}{1-a_{n}}<2(1+L)
$$

Let $M=2(1+L)>1$, (2.1) becomes

$$
\begin{equation*}
\left\|T^{n} x_{n}-x_{n}\right\| \leqslant M\left\|x_{n}-p\right\| \tag{2.2}
\end{equation*}
$$

Using similar procedure we can easily get that

$$
\begin{equation*}
\left\|T^{n} x_{n+1}-x_{n+1}\right\| \leqslant M\left\|x_{n+1}-p\right\| \tag{2.3}
\end{equation*}
$$

This completes the proof.
Lemma 2.1. Mogbademu [9] Let $E$ be a normed linear space then for all $x, y \in E$ and for all $j(x+y) \in J(x+y)$, the following inequality holds:

$$
\|x+y\|^{2} \leqslant\|x\|^{2}+2\langle y, j(x+y)\rangle
$$

Lemma 2.2. C. Moore and B.V Nnoli [10] Let $\Phi:[0,+\infty) \rightarrow[0,+\infty)$ be a strictly increasing function with $\Phi(0)=0$ and let $\left\{\theta_{n}\right\},\left\{\lambda_{n}\right\},\left\{\sigma_{n}\right\}$ be any nonnegative real sequences such that $\sigma_{n}=o\left(\lambda_{n}\right), \sum_{i=0}^{\infty} \lambda_{n}=\infty$, and $\lim _{n \rightarrow \infty} \lambda_{n}=0$. Suppose that

$$
\theta_{n+1}^{2} \leqslant \theta_{n}^{2}-\lambda_{n} \Phi\left(\theta_{n+1}\right)+\sigma_{n}, \quad n \geqslant 1
$$

then $\lim _{n \rightarrow \infty} \theta_{n}=0$.

## 3. The Main Result

Theorem 3.1. Let $E$ be a real Banch space, $K$ be a nonempty closed convex subset of $E, T: K \rightarrow K$ be a Uniformly L-Lipschitzian mapping of Gregus Type with $F(T) \neq \emptyset$ where $F(T)=\{x \in K: T x=x\}$ and $p \in F(T)$. Let $\left\{k_{n}\right\} \in[0, \infty)$ be a sequence of real numbers such that $k_{n} \rightarrow 1$ as $n \rightarrow \infty$, and let $\alpha_{n} \in[0,1]$ satisfying the following:
i) $\alpha_{n} \rightarrow 0$ as $n \rightarrow \infty$
ii) $\sum_{n=0}^{\infty} \alpha_{n}=\infty$;

For any $x_{0} \in K$, define a sequence $\left\{x_{n}\right\}$ by

$$
x_{n+1}=\left(1-\alpha_{n}\right) x_{n}+\alpha_{n} T^{n} x_{n}
$$

for all $n \geqslant 0$. If there exists a strictly increasing function $\Phi:[0,+\infty) \rightarrow[0,+\infty)$ with $\Phi(0)=0$ such that

$$
\begin{equation*}
\left\langle T^{n} x-p, j(x-p)\right\rangle \leqslant k_{n}\|x-p\|^{2}-\Phi(\|x-p\|) \tag{3.1}
\end{equation*}
$$

for all $x \in K$. Then
i) $\left\{x_{n}\right\}_{n \geqslant 0}$ is bounded;
ii) $x_{n} \rightarrow p$ as $n \rightarrow \infty$ where $p$ is a unique fixed point of $T$.

Proof. This proof shall be divided into two steps. In step 1, we will show boundedness, while in step 2 we will show that the iterative sequence $\left\{x_{n}\right\}_{n=0}^{\infty}$ converges strongly to the unique fixed point of $T$ say $p$.

Step 1: Let $k=\sup \left\{k_{n}: n \geqslant 1\right\}$, since $T$ is Uniformly L-Lipschitzian of Gregus Type and $\Phi:[0, \infty) \rightarrow[0, \infty)$ is a strictly increasing real valued function with $\Phi(0)=0$, for $x \in K, p \in F(T)$ we obtain

$$
\Phi(\|x-p\|) \leqslant\left(k+L+\alpha_{n} M\right)\|x-p\|^{2}
$$

Taking limit as $n \rightarrow \infty$ we have

$$
\Phi(\|x-p\|) \leqslant(k+L)\|x-p\|^{2}
$$

Assume that $x_{1} \neq T x_{1}$ for some $x_{1} \in K$ such that

$$
(k+L)\left\|x_{1}-p\right\|^{2} \in R(\Phi)
$$

we denote that $a_{0}=(k+L)\left\|x_{1}-p\right\|^{2}$, where $R(\Phi)$ is the range of $\Phi$. Indeed, if $\Phi(a) \rightarrow \infty$ as $a \rightarrow \infty$, then $a_{0} \in R(\Phi)$; if $\sup \{\Phi(a): a \in[0, \infty)\}=a_{1}<+\infty$ with $a_{1}<a_{0}$, then for $p \in K$, there exists a sequence $\left\{u_{n}\right\}$ in $K$ such that $u_{n} \rightarrow p$ as $n \rightarrow \infty$ with $u_{n} \neq p$, thus there exists an $n_{0} \in \mathbb{N}$ such that

$$
(k+L)\left\|u_{n}-p\right\|^{2}<\frac{a_{1}}{2}
$$

for $n \geqslant n_{0}$. We redefine $x_{1}=u_{n_{0}}$ and $(k+L)\left\|x_{1}-p\right\|^{2} \in R(\Phi)$.
Set $R=\Phi^{-1}\left(a_{0}\right)$. Then we obtain $\left\|x_{1}-p\right\| \leqslant R$.
Denote

$$
B_{1}=\{x \in K:\|x-p\| \leqslant R\}, B_{2}=\{x \in K:\|x-p\| \leqslant 2 R\}
$$

Now, we show that $x_{n} \in B_{1}$ using mathematical induction for any $n \geqslant 1$. If $n=1$, then $x_{1} \in B_{1}$. Suppose that thr result is true for some $n$, that is $x_{n} \in B_{1}$. Now we show that $x_{n+1} \in B_{1}$. Suppose that, $x_{n+1} \notin B_{1}$, that is, $x_{n+1}>R$. Denote

$$
\begin{equation*}
\tau_{0}=\min \left\{1, \frac{1}{2 M}, \frac{1}{2 L M}, \frac{\Phi(R)}{16 R(M(l+3) R}, \frac{\Phi(R)}{16 R^{2}}\right\} \tag{3.2}
\end{equation*}
$$

Since $a_{n}, b_{n}, \alpha_{n}$ and $k_{n}-1 \rightarrow 0$ as $n \rightarrow \infty$. We can let $0 \leqslant a_{n}, b_{n}, \alpha_{n}, k_{n}-1 \leqslant \tau_{0}$ for any $n \geqslant 1$. We obtain the following:

$$
\begin{align*}
\left\|x_{n+1}-p\right\| & \leqslant\left\|x_{n}-p\right\|+\alpha_{n}\left(T^{n} x_{n}-x_{n}\right) \\
& \leqslant R+\alpha_{n} M R \\
& \leqslant 2 R \tag{3.3}
\end{align*}
$$

Using Proposition 2.1 we have the following,

$$
\begin{align*}
\left\|T^{n} x_{n}-T^{n} x_{n+1}\right\| & \leqslant L \alpha_{n}\left\|x_{n}-x_{n+1}\right\|+a_{n}\left\|x_{n}-T^{n} x_{n}\right\|+b_{n}\left\|x_{n+1}-T^{n} x_{n+1}\right\| \\
& \leqslant\left(L \alpha_{n}+a_{n}\right) M\left\|x_{n}-p\right\|+b_{n} M\left\|x_{n+1}-p\right\| \\
& \leqslant\left(L \alpha_{n}+a_{n}\right) M R+2 b_{n} M R \\
& \leqslant \tau_{0} M R(L+3) \\
& \leqslant \frac{\Phi(R)}{16 R} . \tag{3.4}
\end{align*}
$$

Let us consider the following estimate, using Lemma 2.1

$$
\begin{align*}
\left\|x_{n+1}-p\right\|^{2}= & \left\|\left(1-\alpha_{n}\right) x_{n}+\alpha_{n} T^{n} x_{n}-p\right\|^{2} \\
= & \left\|\left(1-\alpha_{n}\right)\left(x_{n}-p\right)+\alpha_{n}\left(T^{n} x_{n}-p\right)\right\|^{2} . \\
\leqslant & \left(1-\alpha_{n}\right)^{2}\left\|x_{n}-p\right\|^{2}+2 \alpha_{n}\left\langle T^{n} x_{n}-p, j\left(x_{n+1}-p\right)\right\rangle \\
= & \left(1-\alpha_{n}\right)^{2}\left\|x_{n}-p\right\|^{2}+2 \alpha_{n}\left\langle T^{n} x_{n}-T^{n} x_{n+1}, j\left(x_{n+1}-p\right)\right\rangle \\
& +2 \alpha_{n}\left\langle T^{n} x_{n+1}-p, j\left(x_{n+1}-p\right)\right\rangle \\
\leqslant & \left(1-\alpha_{n}\right)^{2}\left\|x_{n}-p\right\|^{2}+2 \alpha_{n}\left\|T^{n} x_{n}-T^{n} x_{n+1}\right\|\left\|x_{n+1}-p\right\| \\
& +2 \alpha_{n}\left[k_{n}\left\|x_{n+1}-p\right\|^{2}-\Phi\left(\left\|x_{n+1}-p\right\|\right)\right]  \tag{3.5}\\
\leqslant & \left(1-\alpha_{n}\right)^{2} R^{2}+2 \alpha_{n} k_{n}\left\|x_{n+1}-p\right\|^{2}+ \\
& 2 \alpha_{n} \frac{\Phi(R)}{16 R} 2 R-2 \alpha_{n} \Phi(R) . \tag{3.6}
\end{align*}
$$

Since $\alpha_{n}, k_{n} \rightarrow 1$ as $n \rightarrow \infty$. Thus, without loss of generality, we let $1-2 \alpha_{n} k_{n}>0$ for any $n \geqslant 1$, since

$$
\begin{equation*}
\frac{1}{1-2 \alpha_{n} k_{n}} \leqslant 1+\frac{2 \alpha_{n} k_{n}}{1-2 \alpha_{n} k_{n}} \tag{3.7}
\end{equation*}
$$

from (3.6) we have that

$$
\begin{align*}
\left\|x_{n+1}-p\right\|^{2} \leqslant & R^{2}+\frac{2 \alpha_{n}}{1-2 \alpha_{n} k_{n}}\left[\left(k_{n}-1\right)+\frac{\alpha_{n}}{2}\right] R^{2}+ \\
& \frac{\alpha_{n} \Phi(R)}{4\left(1-2 \alpha_{n} k_{n}\right)}-\frac{2 \alpha_{n}}{1-2 \alpha_{n} k_{n}} \Phi(R) \\
\leqslant & R^{2}+\frac{\alpha_{n} \Phi(R)}{4\left(1-2 \alpha_{n} k_{n}\right)}+\frac{\alpha_{n} \Phi(R)}{4\left(1-2 \alpha_{n} k_{n}\right)}-\frac{2 \alpha_{n}}{1-2 \alpha_{n} k_{n}} \Phi(R) \\
= & R^{2}-\frac{3 \alpha_{n}}{2\left(1-2 \alpha_{n} k_{n}\right)} \Phi(R) \\
\leqslant & R^{2}, \tag{3.8}
\end{align*}
$$

which is a contradiction. Hence, the sequence $\left\{x_{n}\right\}$ is bounded.
Step 2: Here, we intend to show that $x_{n}$ converges uniquely to $p \in F(T)$. Firstly, let us show that $p$ is unique.

Now, we show that $p$ is unique. Suppose for contradiction there exists $p, q \in$ $F(T)$, where $p \neq q$, such that the sequence $\left\{x_{n}\right\}$ converges to $p, q$ hence, we have that,

$$
\begin{aligned}
\|p-q\| & =\left\|T^{n} p-T^{n} q\right\| \\
& \leqslant L\|p-q\|+a_{n}\left\|p-T^{n} p\right\|+b_{n}\left\|q-T^{n} q\right\| \\
& =L\|p-q\|
\end{aligned}
$$

Therefore,

$$
0 \leqslant\|p-q\| \leqslant 0
$$

Hence, $p=q$ is a contradiction.
Next, we show that the sequence $\left\{x_{n}\right\}$ converge to a unique fixed point $p$ of $T$.
Since $\left\|x_{n}-p\right\|$ is bounded, there exists $M_{*}>0$ such that $\left\|x_{n}-p\right\|^{2}<M_{*}$. Hence, from equation (3.5) we have

$$
\begin{align*}
\left\|x_{n+1}-p\right\|^{2} \leqslant & \left(1-\alpha_{n}\right)^{2}\left\|x_{n}-p\right\|^{2}+2 \alpha_{n}\left\|T^{n} x_{n}-T^{n} x_{n+1}\right\|\left\|x_{n+1}-p\right\| \\
& +2 \alpha_{n}\left[k_{n}\left\|x_{n+1}-p\right\|^{2}-\Phi\left(\left\|x_{n+1}-p\right\|\right)\right] \\
\leqslant & \left\|x_{n}-p\right\|^{2}-\alpha_{n} \Phi\left(\| x_{n+1}-p\right) \| \\
& +2 \alpha_{n}\left(k_{n}-1\right) M_{*}^{2}+\alpha_{n}^{2} M_{*}^{2}+2 \alpha_{n} M_{*} M\left(L \alpha_{n}+a_{n}+b_{n}\right) M_{*} \tag{3.9}
\end{align*}
$$

Comparing Lemma 2.2 with (3.9) we can let $\theta_{n+1}^{2}=\left\|x_{n+1}-p\right\|^{2}, \theta_{n}^{2}=\left\|x_{n}-p\right\|^{2}, \lambda_{n}=$ $\alpha_{n}$ and $\sigma_{n}=2 \alpha_{n}\left(k_{n}-1\right) M_{*}^{2}+\alpha_{n}^{2} M_{*}^{2}+2 \alpha_{n} M_{*} M\left(L \alpha_{n}+a_{n}+b_{n}\right) M_{*}$. From condition (i), we have that $\sum_{n=1}^{\infty} \lambda_{n}=\sum_{n=1}^{\infty} \alpha_{n}=\infty$, therefore, $\left\|x_{n}-p\right\| \rightarrow 0$ as $n \rightarrow \infty$.

Example 3.1. Consider $T=\frac{x^{3}}{4+4 x}$, we have shown in Example 1.1, that $T$ is uniformly L-Lipschitzian of Gregus type, and we can easily check that the fixed point of $T$ is $p=0$.

Now, take $\alpha_{n}=1 / 2$, and the initial guess value $x_{0}=0.5,1.0,1.5$ and 2.0. In table 3.1 and figure 3.1, we give a numerical example using Java 2.7, to support our claim that the sequence $x_{n}$ converges uniquely to its fixed point $p=0$ for the uniformly L-Lipschitzian mapping of Gregus type $T$.

Table 3.1: Numerical Example for the uniformly L-Lipschitzian mapping of Gregus type using the iteration in (1.6), $T x=\frac{x^{3}}{4(1+x)}$.

| $S / N$ |  | Modified | Mann Iteration |  |
| :---: | :---: | :---: | :---: | :---: |
| 0 | 0.5 | 1 | 1.5 | 2 |
| 1 | 0.260416667 | 0.5625 | 0.91875 | 1.333333333 |
| 2 | 0.131959801 | 0.295488281 | 0.509897366 | 0.793650794 |
| 3 | 0.066233649 | 0.150233556 | 0.265923844 | 0.43166398 |
| 4 | 0.033150889 | 0.075485267 | 0.134818761 | 0.222854726 |
| 5 | 0.016579852 | 0.037792625 | 0.0676793 | 0.112558723 |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |
| 25 | $1.58 \mathrm{E}-08$ | $3.61 \mathrm{E}-08$ | $6.46 \mathrm{E}-08$ | $1.08 \mathrm{E}-07$ |
| 26 | $7.91 \mathrm{E}-09$ | $1.80 \mathrm{E}-08$ | $3.23 \mathrm{E}-08$ | $5.39 \mathrm{E}-08$ |
| 27 | $3.95 \mathrm{E}-09$ | $9.01 \mathrm{E}-09$ | $1.62 \mathrm{E}-08$ | $2.69 \mathrm{E}-08$ |
| 28 | $1.98 \mathrm{E}-09$ | $4.51 \mathrm{E}-09$ | $8.08 \mathrm{E}-09$ | $1.35 \mathrm{E}-08$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |
| 55 | $1.47 \mathrm{E}-17$ | $3.36 \mathrm{E}-17$ | $6.02 \mathrm{E}-17$ | $1.00 \mathrm{E}-16$ |
| 56 | $7.36 \mathrm{E}-18$ | $1.68 \mathrm{E}-17$ | $3.01 \mathrm{E}-17$ | $5.02 \mathrm{E}-17$ |
| 57 | N/A | $8.40 \mathrm{E}-18$ | $1.50 \mathrm{E}-17$ | $2.51 \mathrm{E}-17$ |
| 58 | N/A | N/A | $7.52 \mathrm{E}-18$ | $1.25 \mathrm{E}-17$ |
| 59 | N/A | N/A | N/A | $6.27 \mathrm{E}-18$ |



Fig. 3.1: Convergence behaviour of Modified Mann iteration process to the fixed point $p=0$ with initial guess values taken at $x_{0}=0.5,1.0,1.5$ and 2.0.

## Competing Interest

The authors declares that they have no competing interest.

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Olilima O. Joshua
Faculty of Science
Department of Mathematical Sciences
Augustine University Ilara-Epe
Lagos, Nigeria
P. M. Box 1010
joshua.olilima@augustineuniversity.edu.ng

Mogbademu A. Adesanmi
Faculty of Science
Department of Mathematical Sciences

Augustine University Ilara-Epe
Lagos, Nigeria
P. M. Box 1010
amogbademu@unilag.edu.ng

Adeniran T. Adefemi
Faculty of Science
Department of Statistics
University of Ibadan
Oyo State, Nigeria
at.adeniran@mail.ui.edu.ng

# ON THE FIXED-CIRCLE PROBLEM 

Ufuk Çelik and Nihal Özgür

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#### Abstract

In this paper, we focus on the geometric properties of fixed-points of a selfmapping and obtain new solutions to a recent problem called "fixed-circle problem" in the setting of an $S$-metric space. For this purpose, we develop various techniques by defining new contractive conditions and using some auxiliary functions. Furthermore, we present new examples to support our theoretical results.


Keywords: fixed-points; $S$-metric space; self-mapping.

## 1. Introduction

It is known that the fixed-point theory has been generalized by various approaches. One of these approaches is to generalize the used contractive condition (for example see [2], [5]). The other is to generalize the used metric space (see $[1,8,21,23]$ and the references therein). For example, in [21], Sedghi, Shobe and Aliouche presented the notion of an $S$-metric space as the generalization of a metric space. Then, some fixed-point theorems have been extensively studied on $S$-metric spaces (see $[6,7,9,13,15,18,19,21,22,24,25,27]$ for more details).

On the other hand, fixed-point theorems have been widely studied with different aspects such as the uniqueness of a fixed-point, common fixed point, etc. If a fixed point is not unique then the investigation of the geometric properties of fixed points of a self-mapping is an interesting problem. As a recent approach, the concept of a fixed circle and the fixed-circle problem have been presented on a metric (resp. an $S$-metric) space as a new direction of the generalization of known fixed-point results (see [17] and [16]). Then, new fixed circle theorems have been given by various techniques on metric (resp. $S$-metric) spaces (see [11, 12, 20, 26] for the metric case; $[10,14,24,25]$ for the $S$-metric case).

Our aim in this paper is to obtain new fixed-circle theorems for self-mappings on an $S$-metric space. In Section 2., we recall some basic facts about $S$-metric spaces.

[^0]In Section 3., we give new fixed-circle theorems by introducing new types of the notion of an $F_{c}^{S}$-contraction introduced and used in [10]. In Section 4., we investigate new existence and uniqueness theorems for fixed circles of self-mappings using some auxiliary functions and contractive conditions. We support our theoretical results by illustrative examples.

## 2. Preliminaries

In this section, we recall some necessary notions and results on $S$-metric spaces with new examples.

Definition 2.1. [21] Let $X$ be a nonempty set and $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ be a function satisfying the following conditions for all $x, y, z, a \in X$ :

1. $\mathcal{S}(x, y, z)=0$ if and only if $x=y=z$,
2. $\mathcal{S}(x, y, z) \leq \mathcal{S}(x, x, a)+\mathcal{S}(y, y, a)+\mathcal{S}(z, z, a)$.

Then $\mathcal{S}$ is called an $S$-metric on $X$ and the pair $(X, \mathcal{S})$ is called an $S$-metric space.
Example 2.1. [21] Let $X=\mathbb{R}$ (or $\mathbb{C}$ ) and the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ be defined by

$$
\mathcal{S}(x, y, z)=|x-z|+|y-z|,
$$

for all $x, y, z \in \mathbb{R}($ or $\mathbb{C})$. Then the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ is an $S$-metric and it is called the usual $S$-metric on $\mathbb{R}$ (or $\mathbb{C}$ ).

Lemma 2.1. [21] Let $(X, \mathcal{S})$ be an $S$-metric space and $x, y \in X$. Then we have

$$
\mathcal{S}(x, x, y)=\mathcal{S}(y, y, x)
$$

It was given the relationships between a metric and an $S$-metric in the following lemma [7].

Lemma 2.2. [7] Let $(X, d)$ be a metric space. Then the following properties are satisfied:

1. $\mathcal{S}_{d}(x, y, z)=d(x, z)+d(y, z)$ for all $x, y, z \in X$ is an $S$-metric on $X$.
2. $x_{n} \rightarrow x$ in $(X, d)$ if and only if $x_{n} \rightarrow x$ in $\left(X, \mathcal{S}_{d}\right)$.
3. $\left\{x_{n}\right\}$ is Cauchy in $(X, d)$ if and only if $\left\{x_{n}\right\}$ is Cauchy in $\left(X, \mathcal{S}_{d}\right)$.
4. $(X, d)$ is complete if and only if $\left(X, \mathcal{S}_{d}\right)$ is complete.

The metric $\mathcal{S}_{d}$ was called as the $S$-metric generated by $d$ in [13].
Now we give a new example of an $S$-metric generated by a metric.

Example 2.2. Let $X \neq \varnothing, d: X^{2} \rightarrow[0, \infty)$ be any metric on $X$ and the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ be defined by

$$
\mathcal{S}(x, y, z)=\min \{1, d(x, z)\}+\min \{1, d(y, z)\}
$$

Then the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ is an $S$-metric on $X$ and the pair $(X, \mathcal{S})$ is an $S$ metric space. Clearly, this $S$-metric $\mathcal{S}$ is generated by the metric $m$ defined as $m(x, y)=$ $\min \{1, d(x, y)\}$.

There are some examples of an $S$-metric which is not generated by any metric (see [7], [10], [14] and [13]). We give a new example.

Example 2.3. Let $X=\mathbb{R}, d: X^{2} \rightarrow[0, \infty)$ be any metric on $X$ and the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ be defined by

$$
\mathcal{S}(x, y, z)=\min \{1, d(x, z)\}+|y-z|
$$

Then the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ is an $S$-metric on $X$ which is not generated by any metric and the pair $(X, \mathcal{S})$ is an $S$-metric space. Conversely, assume that there exists a metric $d_{1}$ such that

$$
\mathcal{S}(x, y, z)=d_{1}(x, z)+d_{1}(y, z),
$$

for all $x, y, z \in X$. Then we obtain

$$
\mathcal{S}(x, x, z)=2 d_{1}(x, z) \Rightarrow d_{1}(x, z)=\frac{1}{2} \min \{1, d(x, z)\}+\frac{1}{2}|x-z|
$$

and

$$
\mathcal{S}(y, y, z)=2 d_{1}(y, z) \Rightarrow d_{1}(y, z)=\frac{1}{2} \min \{1, d(y, z)\}+\frac{1}{2}|y-z|
$$

for all $x, y, z \in X$. So we get

$$
\begin{gathered}
\min \{1, d(x, z)\}+|y-z| \neq \frac{1}{2} \min \{1, d(x, z)\}+\frac{1}{2}|x-z| \\
+\frac{1}{2} \min \{1, d(y, z)\}+\frac{1}{2}|y-z|,
\end{gathered}
$$

which is a contradiction. Hence $\mathcal{S}$ is not generated by any metric.
Definition 2.2. [16] Let $(X, \mathcal{S})$ be an $S$-metric space. Then a circle and a disc are defined on an $S$-metric space as follows, respectively:

$$
C_{x_{0}, r}^{S}=\left\{x \in X: \mathcal{S}\left(x, x, x_{0}\right)=r\right\}
$$

and

$$
D_{x_{0}, r}^{S}=\left\{x \in X: \mathcal{S}\left(x, x, x_{0}\right) \leq r\right\} .
$$

Example 2.4. Let $X$ be a nonempty set, the function $d: X^{2} \rightarrow[0, \infty)$ be any metric on $X$ and the $S$-metric space $(X, \mathcal{S})$ be defined as in Example 2.2. Let us consider the circle $C_{x_{0}, r}^{S}$ according to the $S$-metric $\mathcal{S}$ :

$$
C_{x_{0}, r}^{S}=\left\{x \in X: \mathcal{S}\left(x, x, x_{0}\right)=2 \min \left\{1, d\left(x, x_{0}\right)\right\}=r\right\}
$$

Then we have the following cases:
Case 1: If $r=2$ then $C_{x_{0}, r}^{S}=\left\{x \in X: d\left(x, x_{0}\right) \geq 1\right\}$.
Case 2: If $r>2$ then $C_{x_{0}, r}^{S}=\emptyset$.
Case 3: If $r<2$ then $C_{x_{0}, r}^{S}=C_{x_{0}, \frac{r}{2}}$, where $C_{x_{0}, \frac{r}{2}}=\left\{x \in X: d\left(x, x_{0}\right)=\frac{r}{2}\right\}$.
Example 2.5. Let $X$ be a nonempty set, the function $d: X^{2} \rightarrow[0, \infty)$ be any metric on $X$ and the $S$-metric space be defined as in Example 2.3. Let us consider the circle $C_{x_{0}, r}^{S}$ according to the $S$-metric:

$$
C_{x_{0}, r}^{S}=\left\{x \in X: \mathcal{S}\left(x, x, x_{0}\right)=\min \left\{1, d\left(x, x_{0}\right)\right\}+\left|x-x_{0}\right|=r\right\} .
$$

Then we have the following cases:
Case 1 : If $x \in\left(X \backslash D_{x_{0}, 1}\right) \cup C_{x_{0}, 1}$ then $C_{x_{0}, r}^{S}=\left\{x \in\left(X \backslash D_{x_{0}, 1}\right) \cup C_{x_{0}, 1}:\left|x-x_{0}\right|=r-1\right\}$.
Case 2 : If $x \in D_{x_{0}, 1} \backslash C_{x_{0}, 1}$ then $C_{x_{0}, r}^{S}=\left\{x \in D_{x_{0}, 1} \backslash C_{x_{0}, 1}: d\left(x, x_{0}\right)+\left|x-x_{0}\right|=r\right\}$.
In the following example, the $S$-metric is not generated by any metric but any circle on this $S$-metric space is the same as the circle on the usual metric space $\mathbb{R}$ (or $\mathbb{C})$.

Example 2.6. Let $X=\mathbb{R}$ (or $\mathbb{C})$ and the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ be defined by

$$
\mathcal{S}(x, y, z)=\max \{|x-y|,|y-z|,|z-x|\},
$$

for all $x, y, z \in X$. Then the function $\mathcal{S}: X^{3} \rightarrow[0, \infty)$ is an $S$-metric on $X$ which is not generated by any metric. For any circle $C_{x_{0}, r}^{S}$ on this $S$-metric space we have $C_{x_{0}, r}^{S}=$ $\left\{x_{0}-r, x_{0}+r\right\}$ which is correspond to the circle $C_{x_{0}, r}$ with the equation $\left|y-x_{0}\right|=r$ on the usual metric space $\mathbb{R}$.

## 3. Fixed-Circle Theorems via New Types of $F_{c}^{S}$-contractions

In this section, we give new fixed-circle theorems using new types of the notion of an $F_{c}^{S}$-contraction introduced in [10]. At first, we recall the definition of a fixedcircle and the following family of functions which was introduced by Wardowski in [28].

Definition 3.1. [16] Let $(X, \mathcal{S})$ be an $S$-metric space, $C_{x_{0}, r}^{S}$ be a circle on $X$ and $T: X \rightarrow X$ be a self-mapping. If $T x=x$ for every $x \in C_{x_{0}, r}^{S}$ then the circle $C_{x_{0}, r}^{S}$ is called as the fixed circle of $T$.

Definition 3.2. [28] Let $\mathbb{F}$ be the family of all functions $F:(0, \infty) \rightarrow \mathbb{R}$ such that $(F 1) F$ is strictly increasing,
(F2) For each sequence $\left\{\alpha_{n}\right\}$ in $(0, \infty)$ the following holds $\lim \alpha_{n}=0$ if and only if $\lim F\left(\alpha_{n}\right)=-\infty$,
(F3) There exists $k \in(0,1)$ such that $\lim _{\alpha \rightarrow 0^{+}} \alpha^{k} F(\alpha)=0$.

Some functions that satisfy the conditions $(F 1),(F 2)$ and $(F 3)$ of Definition 3.2 are given in the following example (see [28] for more details).

Example 3.1. [28] The following functions defined by

$$
\begin{gathered}
F_{1}:(0, \infty) \rightarrow \mathbb{R}, F_{1}(x)=\ln (x), \\
F_{2}:(0, \infty) \rightarrow \mathbb{R}, F_{2}(x)=\ln (x)+x, \\
F_{3}:(0, \infty) \rightarrow \mathbb{R}, F_{3}(x)=-\frac{1}{\sqrt{x}}
\end{gathered}
$$

and

$$
F_{4}:(0, \infty) \rightarrow \mathbb{R}, F_{4}(x)=\ln \left(x^{2}+x\right)
$$

are the examples of Definition 3.2.
Using this family of functions, in [4], some new fixed-point theorems was obtained on $S$-metric spaces. In [10], it was introduced the following new contraction type to obtain some fixed-circle results on an $S$-metric space.

Definition 3.3. [10] Let $(X, \mathcal{S})$ be an $S$-metric space. A self-mapping $T$ on $X$ is said to be an $F_{c}^{S}$-contraction if there exist $F \in \mathbb{F}, t>0$ and $x_{0} \in X$ such that for all $x \in X$ the following holds:

$$
\mathcal{S}(T x, T x, x)>0 \Longrightarrow t+F(\mathcal{S}(T x, T x, x)) \leq F\left(\mathcal{S}\left(x, x, x_{0}\right)\right)
$$

In [24], Suzuki-Berinde type $F_{c}^{S}$-contractions were introduced for the same purpose. Now we define new types of $F_{c}^{S}$-contractions to get new fixed-circle results. To do this, we use some classical contraction conditions such as Ćirić-type, modified Hardy-Rogers type and Khan-type contractive conditions.

Let $(X, \mathcal{S})$ be an $S$-metric space and $T$ be a self-mapping on $X$. We will use the number $r$ defined by

$$
\begin{equation*}
r=\inf \{\mathcal{S}(T x, T x, x): x \in X, x \neq T x\} \tag{3.1}
\end{equation*}
$$

in all of our results.

## 3.1. Ćirić type fixed-circle results on $S$-metric spaces

At first, we introduce the following Ćirić type $F_{c}^{S}$-contraction.
Definition 3.4. Let $(X, \mathcal{S})$ be an $S$-metric space and $T$ be a self-mapping on $X$. If there exist $F \in \mathbb{F}, t>0$ and $x_{0} \in X$ such that for all $x \in X$ the following holds:

$$
\mathcal{S}(T x, T x, x)>0 \Longrightarrow t+F(\mathcal{S}(T x, T x, x)) \leq F\left(m\left(x, x, x_{0}\right)\right),
$$

where

$$
m(x, x, y)=\max \left\{\begin{array}{c}
\mathcal{S}(x, x, y), \mathcal{S}(x, x, T x), \mathcal{S}(y, y, T y) \\
\frac{1}{2}[\mathcal{S}(x, x, T y)+\mathcal{S}(y, y, T x)]
\end{array}\right\}
$$

then the self-mapping $T$ is called a Ćirić type $F_{c}^{S}$-contraction on $X$.
An immediate consequence of this definition is the following proposition.
Proposition 3.1. Let $(X, \mathcal{S})$ be an $S$-metric space. If a self-mapping $T$ on $X$ is a Ćirić-type $F_{c}^{S}$-contraction with $x_{0} \in X$ then we have $T x_{0}=x_{0}$.

Proof. Assume that $T x_{0} \neq x_{0}$. From the definition of a Ćirić-type $F_{c}^{S}$-contraction and Lemma 2.1, we get

$$
\begin{aligned}
\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) & >0 \Longrightarrow t+F\left[\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right] \leq F\left(m\left(x_{0}, x_{0}, x_{0}\right)\right) \\
& =F\left(\max \left\{\begin{array}{c}
\mathcal{S}\left(x_{0}, x_{0}, x_{0}\right), \mathcal{S}\left(x_{0}, x_{0}, T x_{0}\right), \mathcal{S}\left(x_{0}, x_{0}, T x_{0}\right), \\
\frac{1}{2}\left[\mathcal{S}\left(x_{0}, x_{0}, T x_{0}\right)+\mathcal{S}\left(x_{0}, x_{0}, T x_{0}\right)\right]
\end{array}\right\}\right) \\
& =F\left(\mathcal{S}\left(x_{0}, x_{0}, T x_{0}\right)\right) .
\end{aligned}
$$

This is a contradiction by the fact that $t>0$. Then we have $T x_{0}=x_{0}$.
Using Ćirić type $F_{c}^{S}$-contractions, we give the following fixed-circle theorem.
Theorem 3.1. Let $(X, \mathcal{S})$ be an $S$-metric space, $T$ be a Ćirić type $F_{c}^{S}$-contractive self-mapping with $x_{0} \in X$ and $r$ be defined as in (3.1). If $\mathcal{S}\left(T x, T x, x_{0}\right)=r$ for all $x \in C_{x_{0}, r}^{S}$ then the circle $C_{x_{0}, r}^{S}$ is a fixed circle of $T$. In particular, $T$ fixes every circle $C_{x_{0}, \rho}^{S}$ where $\rho<r$ if $\mathcal{S}\left(T x, T x, x_{0}\right)=\rho$ for all $x \in C_{x_{0}, \rho}^{S}$.

Proof. Since $\mathcal{S}\left(T x, T x, x_{0}\right)=r$, the self-mapping $T$ maps $C_{x_{0}, r}^{S}$ into (or onto) itself. Let $x \in C_{x_{0}, r}^{S}$ be an arbitrary point. If $T x \neq x$, by the definition of $r$ we have $\mathcal{S}(T x, T x, x) \geq r$. Hence, using the Ćirić-type $F_{c}^{S}$-contractive property, Lemma 2.1, Proposition 3.1 and the fact that $F$ is increasing, we get

$$
\begin{aligned}
F(r) & \leq F(\mathcal{S}(T x, T x, x)) \leq F\left(m\left(x, x, x_{0}\right)\right)-t<F\left(m\left(x, x, x_{0}\right)\right) \\
& =F\left(\max \left\{\begin{array}{c}
\mathcal{S}\left(x, x, x_{0}\right), \mathcal{S}(x, x, T x), \mathcal{S}\left(x_{0}, x_{0}, T x_{0}\right), \\
\left.\frac{1}{2} \mathcal{S}\left(x, x, T x_{0}\right)+\mathcal{S}\left(x_{0}, x_{0}, T x\right)\right]
\end{array}\right\}\right) \\
& =F(\max \{r, \mathcal{S}(x, x, T x), 0, r\})=F(\mathcal{S}(T x, T x, x)),
\end{aligned}
$$

which is a contradiction. Therefore, $\mathcal{S}(T x, T x, x)=0$ and so $T x=x$. Consequently, $C_{x_{0}, r}^{S}$ is a fixed circle of $T$.

Using the similar arguments, it is easy to see that $T$ also fixes any circle $C_{x_{0}, \rho}^{S}$ where $\rho<r$.

Remark 3.1. 1) Notice that, in Theorem 3.1, Ćirić type $F_{c}^{S}$-contractive self-mapping $T$ fixes the disc $D_{x_{0}, r}^{S}$ if $\mathcal{S}\left(T x, T x, x_{0}\right)=\rho$ for all $x \in C_{x_{0}, \rho}^{S}$ and each $\rho \leq r$.
2) In Theorem 3.1, if $r=0$, then we have $C_{x_{0}, r}^{S}=\left\{x_{0}\right\}$ and this is a fixed circle of the self-mapping $T$ by Proposition 3.1.

In the following example, we see that the converse statement of Theorem 3.1 is not always true.

Example 3.2. Let $X=\mathbb{C}$ be the $S$-metric space with the usual $S$-metric defined in Example 2.1, $z_{0} \in \mathbb{C}$ be any point and the self-mapping $T: X \rightarrow X$ be defined as

$$
T z=\left\{\begin{array}{cc}
z & , \quad\left|z-z_{0}\right| \leq \frac{\mu}{2} \\
z_{0} & , \quad\left|z-z_{0}\right|>\frac{\mu}{2}
\end{array}\right.
$$

for all $z \in \mathbb{C}$ with $\mu>0$. We show that $T$ is not a Ćirić-type $F_{c}^{S}$-contractive self-mapping. Indeed, if $\left|z-z_{0}\right|>\frac{\mu}{2}$ for $z \in \mathbb{C}$, then using Lemma 2.1 and the Ćirić-type $F_{c}^{S}$-contractive property, we get

$$
\begin{gathered}
\mathcal{S}(T z, T z, z)=\mathcal{S}\left(z_{0}, z_{0}, z\right)>0 \Longrightarrow t+F\left(\mathcal{S}\left(z_{0}, z_{0}, z\right)\right) \leq F\left(m\left(z, z, z_{0}\right)\right) \\
t+F\left(\mathcal{S}\left(z_{0}, z_{0}, z\right)\right) \leq F\left(\mathcal{S}\left(z, z, z_{0}\right)\right)
\end{gathered}
$$

and so

$$
t+F(r) \leq F(r) \Longrightarrow t \leq 0
$$

This is a contradiction since $t>0$. Hence $T$ is not a Ćirić-type $F_{c}^{S}$-contractive self-mapping for any $z_{0} \in \mathbb{C}$. But $T$ fixes every circle $C_{x_{0}, \rho}^{S}$ where $\rho \leq \mu$.

Now we give some illustrative examples of Theorem 3.1.
Example 3.3. Let $X=\{z \in \mathbb{C}:|z|=2\}$. Let us consider the $S$-metric $\mathcal{S}$ defined in Example 2.6 on $X$ and define the self-mapping $T: X \rightarrow X$ by

$$
T z=\left\{\begin{array}{cc}
-2 & , \quad \frac{\pi}{3} \leq \arg (z) \leq \frac{\pi}{2} \\
z & \text { otherwise }
\end{array}\right.
$$

Then the self-mapping $T$ is a Ćirić-type $F_{c}^{S}$-contractive self-mapping with $F=\ln x, t=$ $\ln \left(\frac{\sqrt{8+4 \sqrt{3}}}{2 \sqrt{3}}\right)$ and $z_{0}=-2 i$. Indeed, we obtain

$$
\begin{aligned}
r & =\inf \{\mathcal{S}(z, z, T z): z \in X, z \neq T z\} \\
& =2 \sqrt{2}
\end{aligned}
$$

In the case $\mathcal{S}(z, z, T z)>0$, we find

$$
\begin{aligned}
m(z, z,-2 i) & =\max \left\{\begin{array}{c}
\mathcal{S}(z, z,-2 i), \mathcal{S}(z, z,-2), \mathcal{S}(-2 i,-2 i,-2 i) \\
\frac{1}{2}[\mathcal{S}(z, z,-2 i)+\mathcal{S}(-2 i,-2 i,-2)]
\end{array}\right\} \\
& =\max \left\{|z+2 i|,|z+2|, 0, \frac{1}{2}[|z+2 i|+|2 i-2|]\right\} \\
& =\sqrt{8+4 \sqrt{3}}
\end{aligned}
$$

and hence

$$
t+\ln (|z+2|) \leq \ln (\sqrt{8+4 \sqrt{3}})
$$

Clearly, $T$ fixes the circle $C_{-2 i, 2 \sqrt{2}}^{S}=\{-2,2\}$ and the disc $D_{-2 i, 2 \sqrt{2}}^{S}=\{z \in X: \mathcal{S}(z, z,-2 i) \leq 2 \sqrt{2}\}$.

### 3.2. Modified Hardy-Rogers type fixed-circle results on $S$-metric spaces

Now we introduce the following modified Hardy-Rogers type $F_{c}^{S}$-contraction.
Definition 3.5. Let $(X, \mathcal{S})$ be an $S$-metric space and $T$ be a self-mapping on $X$. If there exist $F \in \mathbb{F}, t>0$ and $x_{0} \in X$ such that for all $x \in X$ the following holds

$$
\begin{aligned}
\mathcal{S}(T x, T x, x)> & 0 \Longrightarrow t+F(\mathcal{S}(T x, T x, x)) \leq \\
& F\left[\begin{array}{c}
\alpha \mathcal{S}\left(x, x, x_{0}\right)+\beta \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\gamma \mathcal{S}\left(T x, T x, x_{0}\right) \\
+\eta \frac{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)[1+\mathcal{S}(T x, T x, x)]}{\left[1+\mathcal{S}\left(T x_{0}, T x_{0}, x\right)\right]}+\lambda \frac{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)+\mathcal{S}\left(T x, T x, x_{0}\right)}{1+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) \cdot \mathcal{S}\left(x_{0}, x_{0}, x\right)} \\
+\mu+\frac{\mathcal{S}(T x, T x, x)\left[1+\mathcal{S}\left(T x, T x, x_{0}\right)\right]}{1+\mathcal{S}\left(x, x, x_{0}\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}
\end{array}\right]
\end{aligned}
$$

where $\alpha+\beta+\gamma+\eta+\lambda+\mu<\frac{1}{2}, \alpha, \beta, \gamma, \eta, \lambda, \mu \geq 0$ and $a \neq 0$, then the self-mapping $T$ is called a modified Hardy-Rogers type $F_{c}^{S}$-contraction on $X$.

Proposition 3.2. Let $(X, \mathcal{S})$ be an $S$-metric space. If a self-mapping $T$ on $X$ is a modified Hardy-Rogers type $F_{c}^{S}$-contraction with $x_{0} \in X$ then we have $T x_{0}=x_{0}$.

Proof. Assume that $T x_{0} \neq x_{0}$. By the hypothesis, we obtain

$$
\begin{aligned}
\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)> & 0 \Longrightarrow t+F\left(\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right) \leq \\
& F\left[\begin{array}{c}
\alpha \mathcal{S}\left(x_{0}, x_{0}, x_{0}\right)+\beta \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)+\gamma \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) \\
+\eta \frac{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\left[1+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right]}{\left[1+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right]}+\lambda \frac{\left.\left.\mathcal{S} T x_{0}, T x_{0}, x_{0}\right)+\mathcal{S} T x_{0}, T x_{0}, x_{0}\right)}{1+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) \cdot \mathcal{S}\left(x_{0}, x_{0}, x_{0}\right)} \\
+\mu \frac{\left.\mathcal{S} T x_{0}, T x_{0}, x_{0}\right)\left[1+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right]}{1+\mathcal{S}\left(x_{0}, x_{0}, x_{0}\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}
\end{array}\right] \\
= & F\left[(\beta+\gamma+\eta+2 \lambda+\mu) \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right] \\
< & F\left[\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right] .
\end{aligned}
$$

This is a contradiction since $t>0$. Hence we get $T x_{0}=x_{0}$.
Now using the notion of a modified Hardy-Rogers type $F_{c}^{S}$-contraction condition, we prove the following fixed-circle theorem.

Theorem 3.2. Let $(X, \mathcal{S})$ be an $S$-metric space, $T$ be a modified Hardy-Rogers type $F_{c}^{S}$-contractive self-mapping with $x_{0} \in X$ and $r$ be defined as in (3.1). If $\mathcal{S}\left(T x, T x, x_{0}\right)=r$ for all $x \in C_{x_{0}, r}^{S}$ then $C_{x_{0}, r}^{S}$ is a fixed circle of $T$. In particular, $T$ fixes every circle $C_{x_{0}, \rho}^{S}$ where $\rho<r$ if $\mathcal{S}\left(T x, T x, x_{0}\right)=\rho$ for all $x \in C_{x_{0}, \rho}^{S}$.

Proof. Let $x \in C_{x_{0}, r}^{S}$ and $T x \neq x$. If $r=0$, then we have $C_{x_{0}, r}^{S}=\left\{x_{0}\right\}$ and this is a fixed circle of the self-mapping $T$ by Proposition 3.2. Assume that $r>0$. Using the modified Hardy-Rogers type $F_{c}^{S}$-contraction property, Proposition 3.2, Lemma 2.1 and the fact that $F$ is increasing, we get

$$
\begin{aligned}
F(r) & \leq F(\mathcal{S}(T x, T x, x)) \\
& \leq F\left[\begin{array}{c}
\alpha \mathcal{S}\left(x, x, x_{0}\right)+\beta \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\gamma \mathcal{S}\left(T x, T x, x_{0}\right) \\
+\eta \frac{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)[1+\mathcal{S}(T x, T x, x)]}{\left.1+\mathcal{S}\left(T x_{0}, T x_{0}, x\right)\right]}+\lambda \frac{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)+\mathcal{S}\left(T x, T x, x_{0}\right)}{1+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) \cdot \mathcal{S}\left(x, x, x_{0}\right)} \\
+\mu \frac{\mathcal{S}(T x, T x, x)\left[1+\mathcal{S}\left(T x, T x, x_{0}\right)\right]}{1+\mathcal{S}\left(x, x, x_{0}\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}
\end{array}\right]-t \\
& <F[\alpha r+\beta r+\gamma r+\lambda r+\mu \mathcal{S}(T x, T x, x)] \\
& \leq F[(\alpha+\beta+\gamma+\lambda+\mu) \mathcal{S}(T x, T x, x)] \\
& \leq F[\mathcal{S}(T x, T x, x)],
\end{aligned}
$$

which is a contradiction. Therefore, $\mathcal{S}(T x, T x, x)=0$ and so $T x=x$. Consequently, $C_{x_{0}, r}^{S}$ is a fixed circle of $T$. Using the similar arguments, it is easy to see that $T$ also fixes any circle $C_{x_{0}, \rho}^{S}$ where $\rho<r$.

Remark 3.2. 1) Let $(X, S)$ be an $S$-metric space, $T$ be a modified Hardy-Rogers type $F_{c}^{S}$-contractive self-mapping with $x_{0} \in X$ and $r$ be defined as in (3.1). If $\mathcal{S}\left(T x, T x, x_{0}\right)=\rho$ for all $x \in C_{x_{0}, \rho}^{S}$ and each $\rho \leq r$, then $T$ fixes the disc $D_{x_{0}, r}^{S}$.
2) Let us consider the self-mapping $T$ given in Example 3.2. Then it can be easily seen that $T$ is not a modified Hardy-Rogers type $F_{c}^{S}$-contractive self-mapping. But, $T$ fixes every circle $C_{x_{0}, \rho}^{S}$ where $\rho \leq r$. Hence the converse statement of Theorem 3.2 is not always true.

Example 3.4. Let $X=\mathbb{R}^{+}$and the $S$-metric $\mathcal{S}$ be the usual $S$-metric. Let us define the self-mapping $T: X \rightarrow X$ as

$$
T x=\left\{\begin{array}{ccc}
2 x+\frac{4}{x} & , \quad x \in[1,4) \\
x & , & \text { otherwise }
\end{array},\right.
$$

for all $x \in X$. Then the self-mapping $T$ is a modified Hardy-Rogers type $F_{c}^{S}$-contractive self-mapping with $\alpha=\frac{1}{4}, \beta=\frac{1}{25}, \gamma=\frac{1}{25}, \lambda=\frac{1}{25}, \mu=\frac{1}{25}, F=\ln x, t=\ln \frac{9}{8}$ and $x_{0}=35$. Indeed, in the cases $\mathcal{S}(T x, T x, x)>0$ we find

$$
8 \leq \mathcal{S}(T x, T x, x) \leq 10
$$

and

$$
62 \leq \mathcal{S}\left(x, x, x_{0}\right) \leq 68
$$

and hence

$$
\begin{aligned}
t+F\left(2\left|x+\frac{4}{x}\right|\right) & \leq F[2 \alpha|x-35|] \\
& \leq F\left[\begin{array}{c}
2 \alpha|x-35|+2 \beta|x-35|+2 \gamma|T x-35| \\
+\eta \cdot 0+2 \lambda|T x-35| \\
+\mu \frac{2\left|x+\frac{4}{x}\right|[1+|T x-35|]}{1+2|x-35|}
\end{array}\right]
\end{aligned}
$$

Also we have

$$
r=\inf \{\mathcal{S}(T x, T x, x): x \neq T x\}=8
$$

Therefore, the self-mapping $T$ fixes the circle $C_{35,8}^{S}=\{31,39\}$ and the disc $D_{35,8}^{S}=$ $\left\{x \in \mathbb{R}^{+}: 31 \leq x \leq 39\right\}$.

### 3.3. Khan-type fixed-circle results on $S$-metric spaces

Now we introduce the following Khan-type $F_{c}^{S}$-contraction.
Definition 3.6. Let $(X, \mathcal{S})$ be an $S$-metric space and $T$ be a self-mapping on $X$. If there exist $F \in \mathbb{F}, t>0$ and $x_{0} \in X$ such that for all $x \in X$ the following holds:

$$
\begin{aligned}
\mathcal{S}(T x, T x, x) & >0 \Longrightarrow t+F(\mathcal{S}(T x, T x, x)) \\
& \leq F\left[h \frac{\mathcal{S}(T x, T x, x) \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x\right) \mathcal{S}\left(T x, T x, x_{0}\right)}{\mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x, T x, x_{0}\right)}\right]
\end{aligned}
$$

where

$$
h \in[0,1), \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x, T x, x_{0}\right) \neq 0
$$

Then the self-mapping $T$ is called Khan-type $F_{c}^{S}$-contraction on $X$.
Proposition 3.3. Let $(X, \mathcal{S})$ be an $S$-metric space. If a self-mapping $T$ on $X$ is a Khan-type $F_{C}^{S}$-contraction with $x_{0} \in X$. Then we have $T x_{0}=x_{0}$.

Proof. Assume that $T x_{0} \neq x_{0}$. By the hypothesis, we have

$$
\begin{aligned}
\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) & >0 \Longrightarrow t+F\left(\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right) \\
& \leq F\left[h \frac{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right) \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}{\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}\right] \\
& =F\left[h \frac{\mathcal{S}^{2}\left(T x_{0}, T x_{0}, x_{0}\right)+\mathcal{S}^{2}\left(T x_{0}, T x_{0}, x_{0}\right)}{2 \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}\right] \\
& =F\left[h \frac{2 \mathcal{S}^{2}\left(T x_{0}, T x_{0}, x_{0}\right)}{2 \mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)}\right] \\
& <F\left[\mathcal{S}\left(T x_{0}, T x_{0}, x_{0}\right)\right]
\end{aligned}
$$

which is contradiction since $t>0$. Then we have $T x_{0}=x_{0}$.
Now using the notion of a Khan-type $F_{c}^{S}$-contraction condition, we prove the following fixed-circle theorem.

Theorem 3.3. Let $(X, \mathcal{S})$ be an $S$-metric space, $T$ be a Khan-type $F_{c}^{S}$-contraction with $x_{0} \in X$ and $r$ be defined as in (3.1). If $\mathcal{S}\left(T x, T x, x_{0}\right)=r$ for all $x \in C_{x_{0}, r}^{S}$ then $C_{x_{0}, r}^{S}$ is a fixed circle of $T$. In particular, $T$ fixes every circle $C_{x_{0}, \rho}^{S}$ with $\rho<r$ if $\mathcal{S}\left(T x, T x, x_{0}\right)=\rho$ for all $x \in C_{x_{0}, \rho}^{S}$.

Proof. Let $x \in C_{x_{0}, r}^{S}$ and $T x \neq x$. If $r=0$, then we have $C_{x_{0}, r}^{S}=\left\{x_{0}\right\}$ and this is a fixed circle of the self-mapping $T$ by Proposition 3.3.

Assume that $r>0$. Using the Khan-type $F_{C}^{S}$-contractive property, Proposition 3.3, Lemma 2.1 and the fact that $F$ is increasing, we get

$$
\begin{aligned}
F(r) & \leq F(\mathcal{S}(T x, T x, x)) \\
& \leq F\left[h \frac{\mathcal{S}(T x, T x, x) \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x\right) \mathcal{S}\left(T x, T x, x_{0}\right)}{\mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x, T x, x_{0}\right)}\right]-t \\
& <F\left[h \frac{\mathcal{S}(T x, T x, x) r+r^{2}}{2 r}\right]=F\left[h \frac{\mathcal{S}(T x, T x, x)+r}{2}\right] \\
& \leq F\left[h \frac{\mathcal{S}(T x, T x, x)+\mathcal{S}(T x, T x, x)}{2}\right]=F[h \mathcal{S}(T x, T x, x)] \\
& <F[\mathcal{S}(T x, T x, x)],
\end{aligned}
$$

which is a contradiction. Therefore we have $\mathcal{S}(T x, T x, x)=0$ and so $T x=x$. Consequently, $C_{x_{0}, r}^{S}$ is a fixed circle of $T$.

By the similar arguments, it is easy to verify that $T$ also fixes any circle $C_{x_{0}, \rho}^{S}$ where $\rho<r$.

Remark 3.3. Notice that, in Theorem 3.3, Khan-type $F_{c}^{S}$-contractive self-mapping $T$ fixes the disc $D_{x_{0}, r}^{S}$ if $\mathcal{S}\left(T x, T x, x_{0}\right)=\rho$ for all $x \in C_{x_{0}, \rho}^{S}$ and each $\rho \leq r$. Therefore, the center of any fixed circle is also fixed by $T$.

Now we give the following illustrative example.
Example 3.5. Let $X=\left\{e^{k}: k \in \mathbb{N}\right\}$ and the $S$-metric be defined as in [14] such that

$$
\mathcal{S}(x, y, z)=\left|\ln \frac{x}{y}\right|+\left|\ln \frac{x y}{z^{2}}\right|
$$

for all $x, y, z \in X$ (see Example 2.6 on page 12 in [14]). Let us define the self-mapping $T: X \rightarrow X$ as

$$
T x=\left\{\begin{array}{cc}
e x^{2} & , \quad x \in\left\{e^{1}, e^{2}, e^{3}, e^{4}, e^{5}, e^{6}, e^{7}\right\} \\
x & \text { otherwise }
\end{array}\right.
$$

for all $x \in X$. Then the self-mapping $T$ is a Khan-type $F_{c}^{S}$-contractive self-mapping with $F=-\frac{1}{\sqrt{x}}, t=\frac{1}{8}-\frac{1}{4 \sqrt{5}}$ and $x_{0}=e^{23}$. Indeed, in the case $\mathcal{S}(T x, T x, x)>0$, we find

$$
\mathcal{S}(T x, T x, x) \in\{4,6,8,10,12,14,16\}
$$

and

$$
20<h \frac{\mathcal{S}(T x, T x, x) \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x\right) \mathcal{S}\left(T x, T x, x_{0}\right)}{\mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x, T x, x_{0}\right)}
$$

where $h=\frac{20}{21}$. Then we have
$t+F(\mathcal{S}(T x, T x, x)) \leq F\left[h \frac{\mathcal{S}(T x, T x, x) \mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x_{0}, T x_{0}, x\right) \mathcal{S}\left(T x, T x, x_{0}\right)}{\mathcal{S}\left(T x_{0}, T x_{0}, x\right)+\mathcal{S}\left(T x, T x, x_{0}\right)}\right]$.

We obtain

$$
r=\inf \{\mathcal{S}(T x, T x, x): x \neq T x\}=4
$$

and therefore, the self-mapping $T$ fixes the circle $C_{e^{23}, 4}^{S}=\left\{e^{21}, e^{25}\right\}$ and the disc $D_{e^{23,4}}^{S}=$ $\left\{e^{21}, e^{22}, e^{23}, e^{24}, e^{25}\right\}$.

## 4. Fixed-Circle Theorems via Auxiliary Functions

In this section, we investigate the existence and uniqueness theorems for fixed circles of self-mappings using some auxiliary functions. Let $r>0$ be any real number. We consider the function $\varphi_{r}: \mathbb{R}^{+} \cup\{0\} \rightarrow \mathbb{R}$ defined as

$$
\varphi_{r}(u)=\left\{\begin{array}{ccc}
u-r & , \quad u>0  \tag{4.1}\\
0 & , & u=0
\end{array}\right.
$$

for all $u \in \mathbb{R}^{+} \cup\{0\}$ [12]. Using the function $\varphi_{r}$ we give the following theorem.
Theorem 4.1. Let $(X, \mathcal{S})$ be an $S$-metric space and $C_{x_{0}, r}^{S}$ be any circle on $X$. Consider the function $\varphi_{r}$ defined in (4.1). If there exists a self-mapping $T: X \rightarrow X$ satisfying the conditions

1. $\mathcal{S}\left(T x, T x, x_{0}\right)=r$ for each $x \in C_{x_{0}, r}^{S}$,
2. $\mathcal{S}(T x, T x, T y)>r$ for each $x, y \in C_{x_{0}, r}^{S}$ and $x \neq y$,
3. $\mathcal{S}(T x, T x, T y) \leq \mathcal{S}(x, x, y)-\varphi_{r}(\mathcal{S}(x, x, T x))$ for each $x, y \in C_{x_{0}, r}^{S}$,
then the circle $C_{x_{0}, r}^{S}$ is a fixed circle of $T$.
Proof. Let $x \in C_{x_{0}, r}^{S}$ be an arbitrary point. By the condition (1), we have $T x \in$ $C_{x_{0}, r}^{S}$ for all $x \in C_{x_{0}, r}^{S}$. Now we prove that $x$ is a fixed point of $T$. On the contrary, let us assume that $T x \neq x$. Taking $y=T x$ and using the condition (2), we find

$$
\begin{equation*}
\mathcal{S}\left(T x, T x, T^{2} x\right)>r \tag{4.2}
\end{equation*}
$$

Using the condition (3), we have

$$
\begin{align*}
\mathcal{S}\left(T x, T x, T^{2} x\right) & \leq \mathcal{S}(x, x, T x)-\varphi_{r}(\mathcal{S}(x, x, T x))  \tag{4.3}\\
& =\mathcal{S}(x, x, T x)-\mathcal{S}(x, x, T x)+r=r
\end{align*}
$$

Combining the inequalities (4.2) and (4.3), we get a contradiction. Hence it should be $T x=x$. Consequently, the circle $C_{x_{0}, r}^{S}$ is a fixed circle of $T$.

Remark 4.1. Notice that the condition (1) in Theorem 4.1 guarantees that $T x$ is on the circle $C_{x_{0}, r}^{S}$ for $x \in C_{x_{0}, r}^{S}$, the condition (2) shows that the distance of the images of any two elements on the circle $C_{x_{0}, r}^{S}$ can not be less than (or equal to) $r$.

Now we give an example of a self-mapping which has a fixed-circle on an $S$-metric space.

Example 4.1. Let $X=\mathbb{R}$ and the metric function $d: X^{2} \rightarrow[0, \infty)$ be defined by

$$
d(x, y)=\left\{\begin{array}{cll}
0 & , & x=y \\
|x|+|y| & , & x \neq y
\end{array}\right.
$$

for all $x, y \in X$. Let us consider the $S$-metric defined in Example 2.2. The circle $C_{\frac{1}{2}, 1}^{S}=$ $\left\{x \in X: \mathcal{S}\left(x, x, \frac{1}{2}\right)=1\right\}=\{0\}$. If we consider the self-mapping $T_{1}: \mathbb{R} \rightarrow \mathbb{R}$ defined by

$$
T_{1} x=\left\{\begin{array}{ll}
4 & , \quad x=\frac{1}{2} \\
0 & ,
\end{array},\right.
$$

for all $x \in \mathbb{R}$ then the self-mapping $T_{4}$ satisfies the conditions of Theorem 4.1 and $T_{4}$ fixes the circle $C_{\frac{1}{2}, 1}^{S}$.

In the following example, we see that the converse statement of Theorem 4.1 is not always true.

Example 4.2. Let $X=\mathbb{C}$ and consider the $S$-metric defined in Example 2.6. Let us consider the circle $C_{0, \frac{1}{3}}^{S}$ and define the self-mapping $T_{2}: \mathbb{C} \rightarrow \mathbb{C}$

$$
T_{2} z=\left\{\begin{array}{ccc}
\frac{1}{9 z} & , & z \neq 0 \\
0 & , & z=0
\end{array}\right.
$$

for all $z \in \mathbb{C}$, where $\bar{z}$ denotes the complex conjugate of the complex number $z$. Clearly, we have $T_{2}\left(C_{0, \frac{1}{3}}^{S}\right)=\left(C_{0, \frac{1}{3}}^{S}\right)$. It can be easily checked that the self mapping $T_{2}$ does not satisfy the condition (2) of Theorem 4.1. But, an easy computation shows that $T_{2}$ fixes the circle $C_{0, \frac{1}{3}}^{S}$.

In the following example we see that the circle need not to be fixed even if $T\left(C_{x_{0}, r}^{S}\right)=C_{x_{0}, r}^{S}$.

Example 4.3. Let $(\mathbb{C}, \mathcal{S})$ be the usual $S$-metric space. Let us consider the circle $C_{0, \frac{1}{8}}^{S}$ and define the self-mapping $T_{3}: \mathbb{C} \rightarrow \mathbb{C}$ as

$$
T_{3} z=\left\{\begin{array}{ccc}
\frac{1}{16 z} & , & z \neq 0 \\
0 & , & z=0
\end{array}\right.
$$

for all $z \in \mathbb{C}$. Then we have $T_{3}\left(C_{0, \frac{1}{8}}^{S}\right)=C_{0, \frac{1}{8}}^{S}$. But the self-mapping $T_{3}$ does not satisfy the conditions (2) and (3) of Theorem 4.1. Clearly, the circle $C_{0, \frac{1}{8}}^{S}$ is not a fixed circle of $T_{3}$ since $T_{3}\left(\frac{i}{4}\right)=-\frac{i}{4}$ and $T_{3}\left(-\frac{i}{4}\right)=\frac{i}{4}$. More precisely, $T_{3}$ fixes only the points $\frac{1}{4}$ and $-\frac{1}{4}$ on the circle $C_{0, \frac{1}{8}}^{S}$.

In the following example we see that a self mapping can be fix more than one circle.

Example 4.4. Let $X=\mathbb{R}$ and $(X, \mathcal{S})$ be the $S$-metric space defined in Example 2.6. Let us consider the circles $C_{0,4}^{S}$ and $C_{6,2}^{S}$ and the self-mapping $T_{4}: \mathbb{R} \rightarrow \mathbb{R}$ as

$$
T_{4} x=\left\{\begin{array}{ccc}
\frac{2 x+4}{x+5} & , & x \in(-\infty, 4) \\
\frac{17 x+56}{24} & , & x \in(4, \infty) \\
4 & , & x=4
\end{array}\right.
$$

for all $x \in \mathbb{R}$. It can be easily checked that the self-mapping $T_{4}$ satisfies the conditions of Theorem 4.1 and that both of the circles $C_{0,4}^{S}$ and $C_{6,2}^{S}$ are the fixed circles of $T_{4}$.

Now we give another existence theorem for fixed circles.
Theorem 4.2. Let $(X, \mathcal{S})$ be an $S$-metric space and $C_{x_{0}, r}^{S}$ be any circle on $X$. Let us define the mapping

$$
\varphi: X \rightarrow[0, \infty), \varphi(x)=\mathcal{S}\left(x, x, x_{0}\right)
$$

for all $x \in X$. If there exists a self-mapping $T: X \rightarrow X$ satisfying

1. $\mathcal{S}(x, x, T x) \leq \max \{\varphi(x), \varphi(T x)\}-r$,
2. $\mathcal{S}\left(T x, T x, x_{0}\right)-h \mathcal{S}(x, x, T x) \leq r$,
for all $x \in C_{x_{0}, r}^{S}$ and $h \in[0,1)$, then $C_{x_{0}, r}^{S}$ is a fixed circle of $T$.
Proof. Let $x \in C_{x_{0}, r}^{S}$. On the contrary, assume that $T x \neq x$. Then we have the following cases:

Case 1. If max $\{\varphi(x), \varphi(T x)\}=\varphi(x)$ then using the condition (1) we have

$$
\mathcal{S}(x, x, T x) \leq \max \{\varphi(x), \varphi(T x)\}-r=\varphi(x)-r=r-r=0
$$

and so $\mathcal{S}(x, x, T x)=0$, a contradiction. Hence we get $T x=x$.
Case 2. If $\max \{\varphi(x), \varphi(T x)\}=\varphi(T x)$ then we obtain

$$
\mathcal{S}(x, x, T x) \leq \max \{\varphi(x), \varphi(T x)\}-r=\varphi(T x)-r
$$

and using the condition (2) we find

$$
\mathcal{S}(x, x, T x) \leq \varphi(T x)-r \leq h \mathcal{S}(x, x, T x)+r-r=h \mathcal{S}(x, x, T x),
$$

a contradiction since $h \in[0,1)$. Hence we get $T x=x$.
Consequently, $C_{x_{0}, r}^{S}$ is a fixed circle of $T$.

Remark 4.2. (1) Notice that the condition (1) in Theorem 4.2 guarantees that $T x$ is not in the interior of the circle $C_{x_{0}, r}^{S}$ for $x \in C_{x_{0}, r}^{S}$. Similarly the condition (2) guarantees that $T x$ is not exterior of the circle $C_{x_{0}, r}^{S}$ for $x \in C_{x_{0}, r}^{S}$. Hence $T x \in C_{x_{0}, r}^{S}$ for each $x \in C_{x_{0}, r}^{S}$.
(2) Notice that the conditions of Theorem 4.2 are satisfied by the self-mapping $T_{2}$.

Now we give the following example.
Example 4.5. Let $X=\mathbb{R}$ be the $S$-metric space with the usual $S$-metric defined in Example 2.1. Let us consider the circle $C_{0,8}^{S}$ and define the self-mapping $T_{5}: \mathbb{R} \rightarrow \mathbb{R}$ as

$$
T_{5} x=\left\{\begin{array}{ccc}
2 & , & x \in\left\{-\frac{8}{\sqrt{3}}, 2\right\} \\
\frac{8 x+16 \sqrt{3}}{\sqrt{3} x+8} & , & x \in \mathbb{R} \backslash\left\{-\frac{8}{\sqrt{3}}, 2\right\}
\end{array},\right.
$$

for all $x \in \mathbb{R}$. Then the self-mapping $T_{5}$ satisfies the conditions (1) and (2) in Theorem 4.2. Hence $C_{0,8}^{S}$ is a fixed circle of $T_{5}$. Notice that $C_{3,2}^{S}$ is another fixed circle of $T_{5}$ and so the number of the fixed circles need not to be unique for a giving self-mapping.

Now, in the following example, we give an example of a self-mapping which satisfies the condition (1) and does not satisfy the condition (2) of Theorem 4.2.

Example 4.6. Let $X=\mathbb{R}$ and the $S$-metric be defined as in Example 2.6. Let us consider the circle $C_{0,6}^{S}$ and define the self-mapping $T_{6}: \mathbb{R} \rightarrow \mathbb{R}$ as

$$
T_{6} x=\left\{\begin{array}{ccc}
\frac{4 x+48 \sqrt{3}}{\sqrt{3} x+3} & , & x \in(-7,7) \\
20 & , & \text { otherwise }
\end{array}\right.
$$

for all $x \in \mathbb{R}$. Then the self-mapping $T_{6}$ satisfies the conditions (1) but does not satisfy the conditions (2) in Theorem 4.2. Consequently $C_{0,6}^{S}$ is not a fixed circle of $T_{6}$.

In the following, we give an example of a self-mapping which satisfies the condition (2) and does not satisfy the condition (1) in Theorem 4.2.

Example 4.7. Let $X=\mathbb{C}$ be the $S$-metric space with the usual $S$-metric defined in Example 2.1. Let us consider the circle $C_{0,12}^{S}$ and define the self-mapping $T_{7}: \mathbb{C} \rightarrow \mathbb{C}$ as

$$
T_{7} z=\left\{\begin{array}{ccc}
\frac{\operatorname{Re}(z)}{2} & \text { if } & \operatorname{Re}(z) \geq 0 \\
-\frac{\operatorname{Re}(z)}{2} & \text { if } & \operatorname{Re}(z)<0
\end{array},\right.
$$

for all $z \in \mathbb{C}$. Then the self-mapping $T_{7}$ satisfies the condition (2) and does not satisfy the condition (1) in Theorem 4.2.

Now we use the following corollaries to obtain a uniqueness theorem for fixed circles of self-mappings.

Corollary 4.1. [22] Let $(X, \mathcal{S})$ be a complete $S$-metric space and $T$ be a selfmapping of $X$, and

$$
\begin{equation*}
S(T x, T x, T y) \leq a S(x, x, y)+b S(T x, T x, x)+c S(T y, T y, y) \tag{4.4}
\end{equation*}
$$

for some $a, b, c \geq 0, a+b+c<1$, and all $x, y \in X$. Then $T$ has a unique fixed point in $X$. Moreover, if $c<\frac{1}{2}$ then $T$ is continuous at the fixed point.

Corollary 4.2. [22] Let $(X, \mathcal{S})$ be a complete $S$-metric space and $T$ be a selfmapping of $X$, and

$$
\begin{equation*}
S(T x, T x, T y) \leq h \max \{S(T x, T x, y), S(T y, T y, x)\} \tag{4.5}
\end{equation*}
$$

for some $h \in\left[0, \frac{1}{3}\right)$ and all $x, y \in X$. Then $T$ has a unique fixed point in $X$. Moreover, $T$ is continuous at the fixed point.

We give the following theorem.

Theorem 4.3. Let $(X, \mathcal{S})$ be an $S$-metric space and $T: X \rightarrow X$ be a self-mapping with the fixed circle $C_{x_{0}, r}^{S}$. If one of the contractive conditions (4.4) or (4.5) is satisfied for all $x \in C_{x_{0}, r}^{S}, y \in X \backslash C_{x_{0}, r}^{S}$ by $T$ then $C_{x_{0}, r}^{S}$ is the unique fixed circle of $T$.

Proof. Assume that there exists two fixed circles $C_{x_{0}, r}^{S}$ and $C_{x_{0}, \rho}^{S}$ of the self-mapping $T$. Let $x \in C_{x_{0}, r}^{S}$ and $y \in C_{x_{0}, \rho}^{S}$ be arbitrary points with $x \neq y$. If the contractive condition (4.4) is satisfied by $T$, then we obtain

$$
\begin{aligned}
S(x, x, y) & =S(T x, T x, T y) \leq a S(x, x, y)+b S(T x, T x, x)+c S(T y, T y, y) \\
& =a S(x, x, y)
\end{aligned}
$$

which is a contradiction since $a+b+c<1$. Hence it should be $x=y$. Consequently $C_{x_{0}, r}^{S}$ is the unique fixed circle of $T$. Similarly, if the contractive condition (4.5) is satisfied by $T$ then we get
$S(x, x, y)=S(T x, T x, T y) \leq h \max \{S(T x, T x, y), S(T y, T y, x)\}=h S(x, x, y)$,
which is a contradiction since $h \in\left[0, \frac{1}{3}\right)$. Hence it should be $x=y$. Consequently $C_{x_{0}, r}^{S}$ is the unique fixed circle of $T$.

Now we consider the identity map $I_{X}: X \rightarrow X$ defined as $I_{X}(x)=x$ for all $x \in X$. We note that the identity map satisfies the conditions of Theorem 4.2 but can not satisfy the condition (2) of Theorem 4.1 everywhen. Therefore, we investigate a condition which excludes the identity map in Theorem 4.2 (resp. Theorem 4.1). For this purpose, we obtain the following theorem.

Theorem 4.4. Let $(X, \mathcal{S})$ be an $S$-metric space, $T: X \rightarrow X$ be a self mapping having a fixed circle $C_{x_{0}, r}^{S}$ and the mapping $\varphi_{r}$ be defined as in (4.1). The selfmapping $T: X \rightarrow X$ satisfies the condition

$$
\begin{equation*}
\mathcal{S}(x, x, T x)<\varphi_{r}(\mathcal{S}(x, x, T x))+r \tag{4.6}
\end{equation*}
$$

for all $x \in X$ if and only if $T=I_{X}$.

Proof. Let $x \in X$ be any point and assume that $T x \neq x$. Then using the inequality (4.6), we get

$$
\mathcal{S}(x, x, T x)<\varphi_{r}(\mathcal{S}(x, x, T x))+r=\mathcal{S}(x, x, T x)-r+r,
$$

which is a contradiction. Hence we have $T x=x$ and $T=I_{X}$.
Conversely, it is clear that the identity map $I_{X}$ satisfies the condition (4.6).

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Ufuk Çelik
Faculty of Arts and Sciences
Department of Mathematics
10145 Balıkesir, Turkey
ufuk.celik@baun.edu.tr
Nihal Özgür
Faculty of Arts and Sciences
Department of Mathematics
10145 Balıkesir, Turkey
nihal@balikesir.edu.tr

# ON PSEUDO-HERMITIAN MAGNETIC CURVES IN SASAKIAN MANIFOLDS 

Şaban Güvenç and Cihan Özgür

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#### Abstract

We define pseudo-Hermitian magnetic curves in Sasakian manifolds endowed with the Tanaka-Webster connection. After we have given a complete classification theorem, we shall construct parametrizations of pseudo-Hermitian magnetic curves in $\mathbb{R}^{2 n+1}(-3)$.


Keywords: magnetic curve; slant curve; Sasakian manifold; the Tanaka-Webster connection.

## 1. Introduction

The study of the motion of a charged particle in a constant and time-independent static magnetic field on a Riemannian surface is known as the Landau-Hall problem [16]. The main problem is to study the movement of a charged particle moving in the Euclidean plane $\mathbb{E}^{2}$. The solution of the Lorentz equation (called also the Newton equation) corresponds to the motion of the particle. The trajectory of a charged particle moving on a Riemannian manifold under the action of the magnetic field is a very interesting problem from a geometric point of view [16].

Let $(N, g)$ be a Riemannian manifold, and $F$ a closed 2-form, $\Phi$ the Lorentz force, which is a (1,1)-type tensor field on $N . F$ is called a magnetic field if it is associated to $\Phi$ by the relation

$$
\begin{equation*}
F(X, Y)=g(\Phi X, Y) \tag{1.1}
\end{equation*}
$$

where $X$ and $Y$ are vector fields on $N$ (see [1], [3] and [8]). Let $\nabla$ be the Riemannian connection on $N$ and consider a differentiable curve $\alpha: I \rightarrow N$, where $I$ denotes an open interval of $\mathbb{R}$. $\alpha$ is said to be a magnetic curve for the magnetic field $F$, if it is a solution of the Lorentz equation given by

$$
\begin{equation*}
\nabla_{\alpha^{\prime}(t)} \alpha^{\prime}(t)=\Phi\left(\alpha^{\prime}(t)\right) \tag{1.2}
\end{equation*}
$$

From the definition of magnetic curves, it is straightforward to see that their speed is constant. Specifically, unit-speed magnetic curves are called normal magnetic curves [9].

In [9], Druţă-Romaniuc, Inoguchi, Munteanu and Nistor studied magnetic curves in a Sasakian manifold. Magnetic curves in cosymplectic manifolds were studied in [10] by the same authors. In [13], 3-dimensional Berger spheres and their magnetic curves were considered by Inoguchi and Munteanu. Magnetic trajectories of an almost contact metric manifold were studied in [14], by Jleli, Munteanu and Nistor. The classification of all uniform magnetic trajectories of a charged particle moving on a surface under the action of a uniform magnetic field was obtained in [19], by Munteanu. Furthermore, normal magnetic curves in para-Kaehler manifolds were researched in [15], by Jleli and Munteanu. In [17], Munteanu and Nistor obtained the complete classification of unit-speed Killing magnetic curves in $\mathbb{S}^{2} \times \mathbb{R}$. Moreover, in [18], they studied magnetic curves on $\mathbb{S}^{2 n+1}$. 3 -dimensional normal para-contact metric manifolds and their magnetic curves of a Killing vector field were investigated in [5], by Calvaruso, Munteanu and Perrone. In [20], the present authors studied slant curves in contact Riemannian 3-manifolds with pseudo-Hermitian proper mean curvature vector field and pseudo-Hermitian harmonic mean curvature vector field for the Tanaka-Webster connection in the tangent and normal bundles, respectively. The second author gave the parametric equations of all normal magnetic curves in the 3-dimensional Heisenberg group in [21]. Recently, the present authors have also considered slant magnetic curves in $S$-manifolds in [11].

These studies motivate us to investigate pseudo-Hermitian magnetic curves in $(2 n+1)$-dimensional Sasakian manifolds endowed with the Tanaka-Webster connection. In Section 2, we summarize the fundamental definitions and properties of Sasakian manifolds and the unique connection, namely the Tanaka-Webster connection. We give the main classification theorems for pseudo-Hermitian magnetic curves in Section 3. We show that a pseudo-Hermitian magnetic curve cannot have osculating order greater than 3. In the last section, after a brief information on $\mathbb{R}^{2 n+1}(-3)$, we obtain the parametric equations of pseudo-Hermitian magnetic curves in $\mathbb{R}^{2 n+1}(-3)$ endowed with the Tanaka-Webster connection.

## 2. Preliminaries

Let $N$ be a $(2 n+1)$-dimensional Riemannian manifold satisfying the following equations

$$
\begin{gather*}
\phi^{2}(X)=-X+\eta(X) \xi, \quad \eta(\xi)=1, \quad \phi(\xi)=0, \quad \eta \circ \phi=0,  \tag{2.1}\\
g(X, \xi)=\eta(X), \quad g(X, Y)=g(\phi X, \phi Y)+\eta(X) \eta(Y), \tag{2.2}
\end{gather*}
$$

for all vector fields $X, Y$ on $N$, where $\phi$ is a (1,1)-type tensor field, $\eta$ is a 1 -form, $\xi$ is a vector field and $g$ is a Riemannian metric on $N$. In this case, $(N, \phi, \xi, \eta, g)$ is said to be an almost contact metric manifold [2]. Moreover, if $d \eta(X, Y)=\Phi(X, Y)$,
where $\Phi(X, Y)=g(X, \phi Y)$ is the fundamental 2-form of the manifold, then $N$ is said to be a contact metric manifold [2].

Furthermore, if we denote the Nijenhuis torsion of $\phi$ by $[\phi, \phi]$, for all $X, Y$ $\in \chi(N)$, the condition given by

$$
[\phi, \phi](X, Y)=-2 d \eta(X, Y) \xi
$$

is called the normality condition of the almost contact metric structure. An almost contact metric manifold turns into a Sasakian manifold if the normality condition is satisfied [2].

From Lie differentiation operator in the characteristic direction $\xi$, the operator $h$ is defined by

$$
h=\frac{1}{2} L_{\xi} \phi .
$$

It is directly found that the structural operator $h$ is symmetric. It also validates the equations below, where we denote the Levi-Civita connection by $\nabla$ :

$$
\begin{equation*}
h \xi=0, \quad h \phi=-\phi h, \quad \nabla_{X} \xi=-\phi X-\phi h X \tag{2.3}
\end{equation*}
$$

(see [2]).
If we denote the Tanaka-Webster connection on $N$ by $\hat{\nabla}([22],[24])$, then we have

$$
\widehat{\nabla}_{X} Y=\nabla_{X} Y+\eta(X) \phi Y+\left(\widehat{\nabla}_{X} \eta\right)(Y) \xi-\eta(Y) \nabla_{X} \xi
$$

for all vector fields $X, Y$ on $N$. By the use of equations (2.3), the Tanaka-Webster connection can be calculated as

$$
\begin{equation*}
\widehat{\nabla}_{X} Y=\nabla_{X} Y+\eta(X) \phi Y+\eta(Y)(\phi X+\phi h X)-g(\phi X+\phi h X, Y) \xi \tag{2.4}
\end{equation*}
$$

The torsion of the Tanaka-Webster connection is

$$
\begin{equation*}
\widehat{T}(X, Y)=2 g(X, \phi Y) \xi+\eta(Y) \phi h X-\eta(X) \phi h Y \tag{2.5}
\end{equation*}
$$

In a Sasakian manifold, from the fact that $h=0$ (see [2]), the equations (2.4) and (2.5) can be rewritten as:

$$
\begin{gather*}
\widehat{\nabla}_{X} Y=\nabla_{X} Y+\eta(X) \phi Y+\eta(Y) \phi X-g(\phi X, Y) \xi  \tag{2.6}\\
\widehat{T}(X, Y)=2 g(X, \phi Y) \xi
\end{gather*}
$$

The following proposition states why the Tanaka-Webster connection is unique:
Proposition 2.1. [23] The Tanaka-Webster connection on a contact Riemannian manifold $N=(N, \phi, \xi, \eta, g)$ is the unique linear connection satisfying the following four conditions:
(a) $\widehat{\nabla} \eta=0, \widehat{\nabla} \xi=0 ;$
(b) $\widehat{\nabla} g=0, \widehat{\nabla} \phi=0$;
(c) $\widehat{T}(X, Y)=-\eta([X, Y]) \xi, \quad \forall X, Y \in D$;
(d) $\widehat{T}(\xi, \phi Y)=-\phi \widehat{T}(\xi, Y), \quad \forall Y \in D$.

## 3. Magnetic Curves with respect to the Tanaka-Webster Connection

Let $(N, \phi, \xi, \eta, g)$ be an $n$-dimensional Riemannian manifold and $\alpha: I \rightarrow N$ a curve parametrized by arc-length. If there exists $g$-orthonormal vector fields $E_{1}, E_{2}, \ldots, E_{r}$ along $\alpha$ such that

$$
\begin{align*}
E_{1} & =\alpha^{\prime}, \\
\widehat{\nabla}_{E_{1}} E_{1} & =\widehat{k}_{1} E_{2}, \\
\widehat{\nabla}_{E_{1}} E_{2} & =-\widehat{k}_{1} E_{1}+\widehat{k}_{2} E_{3},  \tag{3.1}\\
& \cdots \\
\widehat{\nabla}_{E_{1}} E_{r} & =-\widehat{k}_{r-1} E_{r-1},
\end{align*}
$$

then $\alpha$ is called a Frenet curve for $\hat{\nabla}$ of osculating order $r,(1 \leq r \leq n)$. Here $\widehat{k}_{1}, \ldots, \widehat{k}_{r-1}$ are called pseudo-Hermitian curvature functions of $\alpha$ and these functions are positive valued on $I$. A geodesic for $\widehat{\nabla}$ (or pseudo-Hermitian geodesic) is a Frenet curve of osculating order 1 for $\widehat{\nabla}$. If $r=2$ and $\widehat{k}_{1}$ is a constant, then $\alpha$ is called a pseudo-Hermitian circle. A pseudo-Hermitian helix of order $r(r \geq 3)$ is a Frenet curve for $\hat{\nabla}$ of osculating order $r$ with non-zero positive constant pseudoHermitian curvatures $\widehat{k}_{1}, \ldots, \widehat{k}_{r-1}$. If we shortly state pseudo-Hermitian helix, we mean its osculating order is 3 [7].

Let $N=\left(N^{2 n+1}, \phi, \xi, \eta, g\right)$ be a Sasakian manifold endowed with the TanakaWebster connection $\widehat{\nabla}$. Let us denote the fundamental 2 -form of $N$ by $\Omega$. Then, we have

$$
\begin{equation*}
\Omega(X, Y)=g(X, \phi Y) \tag{3.2}
\end{equation*}
$$

(see [2]). From the fact that $N$ is a Sasakian manifold, we have $\Omega=d \eta$. Hence, $d \Omega=0$, i.e., it is closed. Thus, we can define a magnetic field $F_{q}$ on $N$ by

$$
F_{q}(X, Y)=q \Omega(X, Y)
$$

namely the contact magnetic field with strength $q$, where $X, Y \in \chi(N)$ and $q \in \mathbb{R}$ [14]. We will assume that $q \neq 0$ to avoid the absence of the strength of magnetic field (see [4] and [9]).

From (1.1) and (3.2), the Lorentz force $\Phi$ associated to the contact magnetic field $F_{q}$ can be written as

$$
\Phi=-q \phi .
$$

So the Lorentz equation (1.2) is

$$
\begin{equation*}
\nabla_{E_{1}} E_{1}=-q \phi E_{1}, \tag{3.3}
\end{equation*}
$$

where $\alpha: I \rightarrow N$ is a curve with arc-length parameter, $E_{1}=\alpha^{\prime}$ is the tangent vector field and $\nabla$ is the Levi-Civita connection (see [9] and [14]). By the use of equations (2.6) and (3.3), we have

$$
\begin{equation*}
\widehat{\nabla}_{E_{1}} E_{1}=\left[-q+2 \eta\left(E_{1}\right)\right] \phi E_{1} . \tag{3.4}
\end{equation*}
$$

Definition 3.1. Let $\alpha: I \rightarrow N$ be a unit-speed curve in a Sasakian manifold $N=\left(N^{2 n+1}, \phi, \xi, \eta, g\right)$ endowed with the Tanaka-Webster connection $\widehat{\nabla}$. Then it is called a normal magnetic curve with respect to the Tanaka-Webster connection $\widehat{\nabla}$ (or shortly a pseudo-Hermitian magnetic curve) if it satisfies equation (3.4).

If $\eta\left(E_{1}\right)=\cos \theta$ is a constant, then $\alpha$ is called a slant curve [6]. From the definition of pseudo-Hermitian magnetic curves, we have the following direct result as in the Levi-Civita case:

Proposition 3.1. If $\alpha$ is a pseudo-Hermitian magnetic curve in a Sasakian manifold, then it is a slant curve.

Proof. Let $\alpha: I \rightarrow N$ be a pseudo-Hermitian magnetic curve. Then, we find

$$
\begin{aligned}
\frac{d}{d t} g\left(E_{1}, \xi\right) & =g\left(\widehat{\nabla}_{E_{1}} E_{1}, \xi\right)+g\left(E_{1}, \widehat{\nabla}_{E_{1}} \xi\right) \\
& =g\left(\left[-q+2 \eta\left(E_{1}\right)\right] \phi E_{1}, \xi\right) \\
& =0
\end{aligned}
$$

So we obtain

$$
\eta\left(E_{1}\right)=\cos \theta=\text { constant }
$$

which completes the proof.
As a result, we can rewrite equation (3.4) as

$$
\begin{equation*}
\widehat{\nabla}_{E_{1}} E_{1}=(-q+2 \cos \theta) \phi E_{1}, \tag{3.5}
\end{equation*}
$$

where $\theta$ is the contact angle of $\alpha$. Now, we can state the following theorem:
Theorem 3.1. Let $\left(N^{2 n+1}, \phi, \xi, \eta, g\right)$ be a Sasakian manifold endowed with the Tanaka-Webster connection $\widehat{\nabla}$. Then $\alpha: I \rightarrow N$ is a pseudo-Hermitian magnetic curve if and only if it belongs to the following list:
(a) pseudo-Hermitian non-Legendre slant geodesics (including pseudo-Hermitian geodesics as integral curves of $\xi$ );
(b) pseudo-Hermitian Legendre circles with $\widehat{k}_{1}=|q|$ and having the Frenet frame field (for $\widehat{\nabla}$ )

$$
\left\{E_{1},-\operatorname{sgn}(q) \phi E_{1}\right\} ;
$$

(c) pseudo-Hermitian slant helices with

$$
\widehat{k}_{1}=|-q+2 \cos \theta| \sin \theta, \widehat{k}_{2}=|-q+2 \cos \theta| \varepsilon \cos \theta
$$

and having the Frenet frame field (for $\widehat{\nabla}$ )

$$
\left\{E_{1}, \frac{\delta}{\sin \theta} \phi E_{1}, \frac{\varepsilon}{\sin \theta}\left(\xi-\cos \theta E_{1}\right)\right\}
$$

where $\delta=\operatorname{sgn}(-q+2 \cos \theta), \varepsilon=\operatorname{sgn}(\cos \theta)$ and $\cos \theta \neq \frac{q}{2}$.

Proof. Let us assume that $\alpha: I \rightarrow N$ is a normal magnetic curve with respect to $\widehat{\nabla}$. Consequently, equation (3.5) must be validated. Let us assume $\widehat{k}_{1}=0$. Hence, we have $\cos \theta=\frac{q}{2}$ or $\phi E_{1}=0$. If $\cos \theta=\frac{q}{2}$, then $\alpha$ is a pseudo-Hermitian non-Legendre slant geodesic. Otherwise, $\phi E_{1}=0$ gives us $E_{1}= \pm \xi$. Thus, $\alpha$ is a pseudo-Hermitian geodesic as an integral curve of $\pm \xi$. So we have just proved that $\alpha$ belongs to (a) from the list, if the osculating order $r=1$. Now, let $\widehat{k}_{1} \neq 0$. From equation (3.5) and the Frenet equations for $\widehat{\nabla}$, we find

$$
\begin{equation*}
\widehat{\nabla}_{E_{1}} E_{1}=\widehat{k}_{1} E_{2}=(-q+2 \cos \theta) \phi E_{1} \tag{3.6}
\end{equation*}
$$

Since $E_{1}$ is unit, the equation (2.2) gives us

$$
\begin{equation*}
g\left(\phi E_{1}, \phi E_{1}\right)=\sin ^{2} \theta \tag{3.7}
\end{equation*}
$$

By the use of (3.6) and (3.7), we obtain

$$
\begin{equation*}
\widehat{k}_{1}=|-q+2 \cos \theta| \sin \theta \tag{3.8}
\end{equation*}
$$

which is a constant. Let us denote $\delta=\operatorname{sgn}(-q+2 \cos \theta)$. From (3.8), we can write

$$
\begin{equation*}
\phi E_{1}=\delta \sin \theta E_{2} \tag{3.9}
\end{equation*}
$$

Let us assume $\widehat{k}_{2}=0$, that is, $r=2$. From the fact that $\widehat{k}_{1}$ is a constant, $\alpha$ is a pseudo-Hermitian circle. (3.9) gives us

$$
\eta\left(\phi E_{1}\right)=0=\delta \sin \theta \eta\left(E_{2}\right),
$$

which is equivalent to

$$
\eta\left(E_{2}\right)=0
$$

Differentiating this last equation with respect to $\hat{\nabla}$, we obtain

$$
\widehat{\nabla}_{E_{1}} \eta\left(E_{2}\right)=0=g\left(\widehat{\nabla}_{E_{1}} E_{2}, \xi\right)+g\left(E_{2}, \widehat{\nabla}_{E_{1}} \xi\right) .
$$

Since $\widehat{\nabla} \xi=0$ and $r=2$, we have

$$
g\left(-\widehat{k}_{1} E_{1}, \xi\right)=0
$$

that is, $\eta\left(E_{1}\right)=0$. Hence, $\alpha$ is Legendre and $\cos \theta=0$. From equation (3.8), we get $\widehat{k}_{1}=|q|$. In this case, we also obtain $\delta=-\operatorname{sgn}(q)$ and $E_{2}=-\operatorname{sgn}(q) \phi E_{1}$. We have proved that $\alpha$ belongs to (b) from the list, if the osculating order $r=2$. Now, let us assume $\widehat{k}_{2} \neq 0$. If we use $\widehat{\nabla} \phi=0$, we calculate

$$
\begin{equation*}
\widehat{\nabla}_{E_{1}} \phi E_{1}=\widehat{k}_{1} \phi E_{2} . \tag{3.10}
\end{equation*}
$$

From (2.1) and (3.9), we find

$$
\begin{equation*}
\phi^{2} E_{1}=-E_{1}+\cos \theta \xi=\delta \sin \theta \phi E_{2} \tag{3.11}
\end{equation*}
$$

which gives us

$$
\phi E_{2}=\frac{\delta}{\sin \theta}\left(-E_{1}+\cos \theta \xi\right)
$$

So equation (3.10) becomes

$$
\begin{equation*}
\widehat{\nabla}_{E_{1}} \phi E_{1}=\widehat{k}_{1} \frac{\delta}{\sin \theta}\left(-E_{1}+\cos \theta \xi\right) \tag{3.12}
\end{equation*}
$$

If we differentiate the equation (3.9) with respect to $\hat{\nabla}$, we also have

$$
\begin{align*}
\widehat{\nabla}_{E_{1}} \phi E_{1} & =\delta \sin \theta \widehat{\nabla}_{E_{1}} E_{2}  \tag{3.13}\\
& =\delta \sin \theta\left(-\widehat{k}_{1} E_{1}+\widehat{k}_{2} E_{3}\right)
\end{align*}
$$

By the use of (3.12) and (3.13), we obtain

$$
\begin{equation*}
\widehat{k}_{1} \cot \theta\left(\xi-\cos \theta E_{1}\right)=\widehat{k}_{2} \sin \theta E_{3} \tag{3.14}
\end{equation*}
$$

One can easily see that

$$
g\left(\xi-\cos \theta E_{1}, \xi-\cos \theta E_{1}\right)=\sin ^{2} \theta
$$

From (3.14), we calculate

$$
\widehat{k}_{2}=|-q+2 \cos \theta| \varepsilon \cos \theta
$$

where we denote $\varepsilon=\operatorname{sgn}(\cos \theta)$. As a result, we get

$$
\begin{gather*}
E_{3}=\frac{\varepsilon}{\sin \theta}\left(\xi-\cos \theta E_{1}\right)  \tag{3.15}\\
E_{2}=\frac{\delta}{\sin \theta} \phi E_{1}
\end{gather*}
$$

If we differentiate (3.15) with respect to $\widehat{\nabla}$, since $\phi E_{1} \| E_{2}$, we find $\widehat{k}_{3}=0$. So we have just completed the proof of (c). Considering the fact that $\widehat{k}_{3}=0$, the Gram-Schmidt process ends. Thus, the list is complete.

Conversely, let $\alpha: I \rightarrow N$ belong to the given list. It is easy to show that equation (3.5) is satisfied. Hence, $\alpha$ is a pseudo-Hermitian magnetic curve.

A pseudo-Hermitian geodesic is said to be a pseudo-Hermitian $\phi$-curve if the set $s p\left\{E_{1}, \phi E_{1}, \xi\right\}$ is $\phi$-invariant. A Frenet curve of osculating order $r=2$ is said to be a pseudo-Hermitian $\phi$-curve if $s p\left\{E_{1}, E_{2}, \xi\right\}$ is $\phi$-invariant. A Frenet curve of osculating order $r \geq 3$ is said to be a pseudo-Hermitian $\phi$-curve if $\operatorname{sp}\left\{E_{1}, E_{2}, \ldots, E_{r}\right\}$ is $\phi$-invariant.

Theorem 3.2. Let $\alpha: I \rightarrow N$ be a pseudo-Hermitian $\phi$-helix of order $r \leq 3$, where $N=\left(N^{2 n+1}, \phi, \xi, \eta, g\right)$ is a Sasakian manifold endowed with the Tanaka-Webster connection $\hat{\nabla}$. Then:
(a) If $\cos \theta= \pm 1$, then it is an integral curve of $\xi$, i.e. a pseudo-Hermitian geodesic and it is a pseudo-Hermitian magnetic curve for $F_{q}$ for arbitrary $q$;
(b) If $\cos \theta \notin\{-1,0,1\}$ and $\widehat{k}_{1}=0$, then it is a pseudo-Hermitian non-Legendre slant geodesic and it is a pseudo-Hermitian magnetic curve for $F_{2 \cos \theta}$;
(c) If $\cos \theta=0$ and $\widehat{k}_{1} \neq 0$, i.e. $\alpha$ is a Legendre $\phi$-curve, then it is a pseudoHermitian magnetic circle generated by $F_{-\delta \widehat{k}_{1}}$, where $\delta=\operatorname{sgn}\left(g\left(\phi E_{1}, E_{2}\right)\right)$;
(d) If $\cos \theta=\frac{\varepsilon \widehat{k}_{2}}{\sqrt{\widehat{k}_{1}^{2}+\widehat{k}_{2}^{2}}}$ and $\widehat{k}_{2} \neq 0$, then it is a pseudo-Hermitian magnetic curve for $F_{-\delta \sqrt{\widehat{k}_{1}^{2}+\widehat{k}_{2}^{2}}+\frac{2 \varepsilon \widehat{k}_{2}}{\sqrt{\hat{k}_{1}^{2}+\widehat{k}_{2}^{2}}}}$, where $\delta=\operatorname{sgn}\left(g\left(\phi E_{1}, E_{2}\right)\right)$ and $\varepsilon=\operatorname{sgn}(\cos \theta)$.
(e) Except above cases, $\alpha$ cannot be a pseudo-Hermitian magnetic curve for any $F_{q}$.

Proof. Firstly, let us assume $\cos \theta= \pm 1$, that is, $E_{1}= \pm \xi$. As a result, we have

$$
\widehat{\nabla}_{E_{1}} E_{1}=0, \phi E_{1}=0
$$

Hence, equation (3.5) is satisfied for arbitrary $q$. This proves (a). Now, let us take $\cos \theta \notin\{-1,0,1\}$ and $\widehat{k}_{1}=0$. In this case, we obtain

$$
\widehat{\nabla}_{E_{1}} E_{1}=0, \phi E_{1} \neq 0
$$

So equation (3.5) is valid for $q=2 \cos \theta$. The proof of (b) is over. Next, let us assume $\cos \theta=0$ and $\widehat{k}_{1} \neq 0$. One can easily see that $\alpha$ has the Frenet frame field (for $\widehat{\nabla}$ )

$$
\left\{E_{1}, \delta \phi E_{1}\right\}
$$

where $\delta$ corresponds to the sign of $g\left(\phi E_{1}, E_{2}\right)$. Consequently, we get

$$
\widehat{\nabla}_{E_{1}} E_{1}=\delta \widehat{k}_{1} \phi E_{1}
$$

that is, $\alpha$ is a pseudo-Hermitian magnetic curve for $q=-\delta \widehat{k}_{1}$. We have just proven (c). Finally, let $\cos \theta=\frac{\varepsilon \widehat{k}_{2}}{\sqrt{\widehat{k}_{1}^{2}+\widehat{k}_{2}^{2}}}$ and $\widehat{k}_{2} \neq 0$. So $\alpha$ has the Frenet frame field (for $\widehat{\nabla}$ )

$$
\left\{E_{1}, \frac{\delta}{\sin \theta} \phi E_{1}, \frac{\varepsilon}{\sin \theta}\left(\xi-\cos \theta E_{1}\right)\right\}
$$

where $\delta=\operatorname{sgn}\left(g\left(\phi E_{1}, E_{2}\right)\right)$ and $\varepsilon=\operatorname{sgn}(\cos \theta)$. After calculations, it is easy to show that equation (3.5) is satisfied for $q=-\delta \sqrt{\widehat{k}_{1}^{2}+\widehat{k}_{2}^{2}}+\frac{2 \varepsilon \widehat{k}_{2}}{\sqrt{\widehat{k}_{1}^{2}+\widehat{k}_{2}^{2}}}$. Hence, the proof of (d) is completed. Except above cases, from Theorem 3.1, $\alpha$ cannot be a pseudo-Hermitian magnetic curve for any $F_{q}$.

## 4. Parametrizations of pseudo-Hermitian magnetic curves in $\mathbb{R}^{2 n+1}(-3)$

In this section, our aim is to obtain parametrizations of pseudo-Hermitian magnetic curves in $\mathbb{R}^{2 n+1}(-3)$. To do this, we need to recall some notions from [2]. Let $N=\mathbb{R}^{2 n+1}$. Let us denote the coordinate functions of $N$ with $\left(x_{1}, \ldots, x_{n}, y_{1}, \ldots, y_{n}, z\right)$. One may define a structure on $N$ by $\eta=\frac{1}{2}\left(d z-\sum_{i=1}^{n} y_{i} d x_{i}\right)$, which is a contact structure, since $\eta \wedge(d \eta)^{n} \neq 0$. This contact structure has the characteristic vector field $\xi=2 \frac{\partial}{\partial z}$. Let us also consider a $(1,1)$-type tensor field $\phi$ given by the matrix form as

$$
\phi=\left[\begin{array}{ccc}
0 & \delta_{i j} & 0 \\
-\delta_{i j} & 0 & 0 \\
0 & y_{j} & 0
\end{array}\right]
$$

Finally, let us take the Riemannian metric on $N$ given by $g=\eta \otimes \eta+\frac{1}{4} \sum_{i=1}^{n}\left(\left(d x_{i}\right)^{2}+\right.$ $\left.\left(d y_{i}\right)^{2}\right)$. It is known that $(N, \phi, \xi, \eta, g)$ is a Sasakian space form and its $\phi$-sectional curvature is $c=-3$. This special Sasakian space form is denoted by $\mathbb{R}^{2 n+1}(-3)$ [2]. One can easily show that the vector fields

$$
\begin{equation*}
X_{i}=2 \frac{\partial}{\partial y_{i}}, X_{n+i}=\phi X_{i}=2\left(\frac{\partial}{\partial x_{i}}+y_{i} \frac{\partial}{\partial z}\right), i=\overline{1, n}, \xi=2 \frac{\partial}{\partial z} \tag{4.1}
\end{equation*}
$$

are $g$-unit and $g$-orthogonal. Hence, they form a $g$-orthonormal basis [2]. Using this basis, the Levi-Civita connection of $\mathbb{R}^{2 n+1}(-3)$ can be obtained as

$$
\begin{gathered}
\nabla_{X_{i}} X_{j}=\nabla_{X_{m+i}} X_{m+j}=0, \nabla_{X_{i}} X_{m+j}=\delta_{i j} \xi, \nabla_{X_{m+i}} X_{j}=-\delta_{i j} \xi \\
\nabla_{X_{i}} \xi=\nabla_{\xi} X_{i}=-X_{m+i}, \nabla_{X_{m+i}} \xi=\nabla_{\xi} X_{m+i}=X_{i}
\end{gathered}
$$

(see [2]). As a result, the Tanaka-Webster connection of $\mathbb{R}^{2 n+1}(-3)$ is

$$
\begin{aligned}
\widehat{\nabla}_{X_{i}} X_{j} & =\widehat{\nabla}_{X_{m+i}} X_{m+j}=\widehat{\nabla}_{X_{i}} X_{m+j}=\widehat{\nabla}_{X_{m+i}} X_{j}= \\
\widehat{\nabla}_{X_{i}} \xi & =\widehat{\nabla}_{\xi} X_{i}=\widehat{\nabla}_{X_{m+i}} \xi=\widehat{\nabla}_{\xi} X_{m+i}=0
\end{aligned}
$$

which was calculated in [12]. Now, we can investigate the parametric equations of pseudo-Hermitian magnetic curves in $\mathbb{R}^{2 n+1}(-3)$ endowed with the Tanaka-Webster connection.

Let $N=\mathbb{R}^{2 n+1}(-3)$ endowed with the Tanaka-Webster connection $\hat{\nabla}$. Let $\alpha: I \subseteq \mathbb{R} \rightarrow N, \alpha=\left(\alpha_{1}, \alpha_{2}, \ldots, \alpha_{n}, \alpha_{n+1}, \ldots, \alpha_{2 n}, \alpha_{2 n+1}\right)$ be a pseudo-Hermitian magnetic curve. Then, the tangential vector field of $\alpha$ can be written as

$$
E_{1}=\sum_{i=1}^{n} \alpha_{i}^{\prime} \frac{\partial}{\partial x_{i}}+\sum_{i=1}^{n} \alpha_{n+i}^{\prime} \frac{\partial}{\partial y_{i}}+\alpha_{2 n+1}^{\prime} \frac{\partial}{\partial z}
$$

In terms of the $g$-orthonormal basis, $E_{1}$ is rewritten as

$$
E_{1}=\frac{1}{2}\left[\sum_{i=1}^{n} \alpha_{n+i}^{\prime} X_{i}+\sum_{i=1}^{n} \alpha_{i}^{\prime} X_{n+i}+\left(\alpha_{2 n+1}^{\prime}-\sum_{i=1}^{n} \alpha_{i}^{\prime} \alpha_{n+i}\right) \xi\right] .
$$

From Proposition 3.1, $\alpha$ is a slant curve. Hence, we have

$$
\eta\left(E_{1}\right)=\cos \theta=\text { constant }
$$

which is equivalent to

$$
\begin{equation*}
\alpha_{2 n+1}^{\prime}=2 \cos \theta+\sum_{i=1}^{n} \alpha_{i}^{\prime} \alpha_{n+i} . \tag{4.2}
\end{equation*}
$$

From the fact that $\alpha$ is parametrized by arc-length, we also have

$$
g\left(E_{1}, E_{1}\right)=1,
$$

that is,

$$
\begin{equation*}
\sum_{i=1}^{2 n}\left(\alpha_{i}^{\prime}\right)^{2}=4 \sin ^{2} \theta \tag{4.3}
\end{equation*}
$$

Differentiating $E_{1}$ with respect to $\widehat{\nabla}$, we obtain

$$
\widehat{\nabla}_{E_{1}} E_{1}=\frac{1}{2}\left(\sum_{i=1}^{n} \alpha_{n+i}^{\prime \prime} X_{i}+\sum_{i=1}^{n} \alpha_{i}^{\prime \prime} X_{n+i}\right)
$$

We also easily find

$$
\phi E_{1}=\frac{1}{2}\left(-\sum_{i=1}^{n} \alpha_{i}^{\prime} X_{i}+\sum_{i=1}^{n} \alpha_{n+i}^{\prime} X_{n+i}\right)
$$

Since $\alpha$ is a pseudo-Hermitian magnetic curve, it must satisfy

$$
\widehat{\nabla}_{E_{1}} E_{1}=(-q+2 \cos \theta) \phi E_{1}
$$

Then, we can write

$$
\frac{\alpha_{n+1}^{\prime \prime}}{-\alpha_{1}^{\prime}}=\ldots=\frac{\alpha_{2 n}^{\prime \prime}}{-\alpha_{n}^{\prime}}=\frac{\alpha_{1}^{\prime \prime}}{\alpha_{n+1}^{\prime}}=\ldots=\frac{\alpha_{n}^{\prime \prime}}{\alpha_{2 n}^{\prime}}=-\lambda
$$

where $\lambda=q-2 \cos \theta$. From the last equations, we can select the pairs

$$
\begin{equation*}
\frac{\alpha_{n+1}^{\prime \prime}}{-\alpha_{1}^{\prime}}=\frac{\alpha_{1}^{\prime \prime}}{\alpha_{n+1}^{\prime}}, \ldots, \frac{\alpha_{2 n}^{\prime \prime}}{-\alpha_{n}^{\prime}}=\frac{\alpha_{n}^{\prime \prime}}{\alpha_{2 n}^{\prime}} . \tag{4.4}
\end{equation*}
$$

Firstly, let $\lambda \neq 0$. Solving the ODEs, we have

$$
\left(\alpha_{i}^{\prime}\right)^{2}+\left(\alpha_{n+i}^{\prime}\right)^{2}=c_{i}^{2}, i=1, \ldots, n
$$

for some arbitrary constants $c_{i}(i=1, \ldots, n)$ such that

$$
\sum_{i=1}^{n} c_{i}^{2}=4 \sin ^{2} \theta
$$

So we have

$$
\alpha_{i}^{\prime}=c_{i} \cos f_{i}, \alpha_{n+i}^{\prime}=c_{i} \sin f_{i}
$$

for some differentiable functions $f_{i}: I \rightarrow \mathbb{R}(i=1, \ldots, n)$. From (4.4), we get

$$
\frac{\alpha_{n+i}^{\prime \prime}}{-\alpha_{i}^{\prime}}=-f_{i}^{\prime}=-\lambda
$$

which gives us

$$
f_{i}=\lambda t+d_{i}
$$

for some arbitrary constants $d_{i}(i=1, \ldots, n)$. Here, $t$ denotes the arc-length parameter. Then, we find

$$
\alpha_{i}^{\prime}=c_{i} \cos \left(\lambda t+d_{i}\right), \alpha_{n+i}^{\prime}=c_{i} \sin \left(\lambda t+d_{i}\right) .
$$

Finally, we obtain

$$
\begin{gathered}
\alpha_{i}=\frac{c_{i}}{\lambda} \sin \left(\lambda t+d_{i}\right)+h_{i} \\
\alpha_{n+i}=\frac{-c_{i}}{\lambda} \cos \left(\lambda t+d_{i}\right)+h_{n+i} \\
\alpha_{2 n+1}=2 t \cos \theta+\sum_{i=1}^{n}\left\{\frac{-c_{i}^{2}}{4 \lambda^{2}}\left[2\left(\lambda t+d_{i}\right)+\sin \left(2\left(\lambda t+d_{i}\right)\right)\right]\right. \\
\left.+\frac{c_{i} h_{n+i}}{\lambda} \sin \left(\lambda t+d_{i}\right)\right\}+h_{2 n+1}
\end{gathered}
$$

for some arbitrary constants $h_{i}(i=1, \ldots, 2 n+1)$.
Secondly, let $\lambda=0$. In this case, $q=2 \cos \theta$ and $\widehat{k}_{1}=0$. Hence, we have

$$
\widehat{\nabla}_{E_{1}} E_{1}=\frac{1}{2}\left(\sum_{i=1}^{n} \alpha_{n+i}^{\prime \prime} X_{i}+\sum_{i=1}^{n} \alpha_{i}^{\prime \prime} X_{n+i}\right)=0
$$

which gives us

$$
\begin{gathered}
\alpha_{i}=c_{i} t+d_{i}, i=1, \ldots, 2 n \\
\alpha_{2 n+1}=2 t \cos \theta+\sum_{i=1}^{n} c_{i}\left(\frac{c_{n+i}}{2} t^{2}+d_{n+i} t\right)+c_{2 n+1}
\end{gathered}
$$

where $c_{i}(i=1,2, \ldots, 2 n+1)$ and $d_{i}(i=1,2, \ldots, 2 n)$ are arbitrary constants such that

$$
\sum_{i=1}^{2 n} c_{i}^{2}=4 \sin ^{2} \theta
$$

To conclude, we can state the following theorem:

Theorem 4.1. The pseudo-Hermitian magnetic curves on $\mathbb{R}^{2 n+1}(-3)$ endowed with the Tanaka-Webster connection have the parametric equations

$$
\alpha: I \subseteq \mathbb{R} \rightarrow \mathbb{R}^{2 n+1}(-3), \alpha=\left(\alpha_{1}, \alpha_{2}, \ldots, \alpha_{n}, \alpha_{n+1}, \ldots, \alpha_{2 n}, \alpha_{2 n+1}\right)
$$

where $\alpha_{i}(i=1, \ldots, 2 n+1)$ satisfies either
(a)

$$
\begin{gathered}
\alpha_{i}=\frac{c_{i}}{\lambda} \sin \left(\lambda t+d_{i}\right)+h_{i} \\
\alpha_{n+i}=\frac{-c_{i}}{\lambda} \cos \left(\lambda t+d_{i}\right)+h_{n+i} \\
\alpha_{2 n+1}=2 \cos \theta t+\sum_{i=1}^{n}\left\{\frac{-c_{i}^{2}}{4 \lambda^{2}}\left[2\left(\lambda t+d_{i}\right)+\sin \left(2\left(\lambda t+d_{i}\right)\right)\right]\right. \\
\left.+\frac{c_{i} h_{n+i}}{\lambda} \sin \left(\lambda t+d_{i}\right)\right\}+h_{2 n+1},
\end{gathered}
$$

where $\lambda=q-2 \cos \theta \neq 0, c_{i}, d_{i}(i=1, \ldots, n)$ and $h_{i}(i=1, \ldots, 2 n+1)$ are arbitrary constants such that

$$
\sum_{i=1}^{n} c_{i}^{2}=4 \sin ^{2} \theta
$$

or
(b)

$$
\begin{gathered}
\alpha_{i}=c_{i} t+d_{i} \\
\alpha_{2 n+1}=2 t \cos \theta+\sum_{i=1}^{n} c_{i}\left(\frac{c_{n+i}}{2} t^{2}+d_{n+i} t\right)+c_{2 n+1}
\end{gathered}
$$

where $q=2 \cos \theta$ and $c_{i}(i=1,2, \ldots, 2 n+1), d_{i}(i=1,2, \ldots, 2 n)$ are arbitrary constants such that

$$
q^{2}+\sum_{i=1}^{2 n} c_{i}^{2}=4
$$

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Şaban Güvenç
Faculty of Arts and Sciences
Department of Mathematics
Campus of Çağıs
10145 Balikesir, Turkey
sguvenc@balikesir.edu.tr

Cihan Özgür
Faculty of Arts and Sciences
Department of Mathematics
Campus of Çağıs
10145 Balikesir, Turkey
cozgur@balikesir.edu.tr

# SOME RESULTS ON $*-$ RICCI FLOW 

Dipankar Debnath and Nirabhra Basu

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Abstract. In this paper we have introduced the notion of $*$-Ricci flow and shown that *-Ricci soliton which was introduced by Kaimakamis and Panagiotidou in 2014 is a self similar soliton of the $*$-Ricci flow. We have also found the deformation of geometric curvature tensors under $*$-Ricci flow. In the last two section of the paper, we have found the $\mathfrak{F}$-functional and $\omega$-functional for $*$-Ricci flow respectively.
Keywords: $*-$ Ricci flow, Conformal Ricci flow, $\mathfrak{F}$ functionals, $\omega$ functionals.

## 1. Introduction

A Ricci soliton is a natural generalization of an Einstein metric and is defined on a Riemannian manifold $(M, g)$ by

$$
\begin{equation*}
£_{V} g+2 S+2 \lambda g=0 \tag{1.1}
\end{equation*}
$$

where $£$ denotes the Lie derivative operator, $\lambda$ is a constant and $S$ is the Ricci tensor of the metric $g$. Tachibana [3] first introduced $*$-Ricci tensor on almost Hermitian manifolds and Hamada [1] apply this to almost contact manifolds by defining

$$
S^{*}(X, Y)=\frac{1}{2} \operatorname{trace}(Z \rightarrow R(X, \phi Y) \phi Z)
$$

for any $X, Y \in T M$. In 2014, Kaimakamis and Panagiotidou [2] introduced the concept of $*$-Ricci solitons within the background of real hypersurfaces of a complex space form, where they essentially modified the definition of Ricci soliton by replacing the Ricci tensor $S$ in (1.1) with the $*$-Ricci tensor $S^{*}$. More precisely, a *-Ricci soliton on $(M, g)$ is defined by

$$
\begin{equation*}
£_{V} g+2 S^{*}+2 \lambda g=0 \tag{1.2}
\end{equation*}
$$

Inspired by the work of Kaimakamis and Panagiotidou [2], we introduced and studied $*$-Ricci flow on Riemannian manifold and further studied $*$-Ricci solitons. We

[^1]have obtained deformation of geometric curvature tensor under $*$-Ricci flow. We have also provided the rate of change of $F$-functionals and $\omega$-entropy functional with respect to time under this flow.
We have defined $*$-Ricci flow as follows
\[

$$
\begin{equation*}
\frac{\partial g}{\partial t}=-2 S^{*}(X, Y) \tag{1.3}
\end{equation*}
$$

\]

In this paper we have shown that just like Ricci soliton; $*$-Ricci soliton is a selfsimilar soliton of the $*$-Ricci flow. We have also found the deformation of geometric curvature tensors under $*$-Ricci flow.

Proposition 1.1. Defining $g \overline{(t)}=\sigma(t) \phi_{t}^{*}(g)+\sigma(t) \phi_{t}^{*}\left(\frac{\partial g}{\partial t}\right)+\sigma(t) \varphi_{t}^{*}\left(£_{X} g\right)$, we have

$$
\begin{equation*}
\frac{\partial \hat{g}}{\partial t}=\dot{\sigma}(t) \psi_{t}^{*}(g)+\sigma(t)+\psi_{t}^{*}\left(\frac{\partial g}{\partial t}\right)+\sigma(t) \psi_{t}^{*}\left(£_{X} g\right) \tag{1.4}
\end{equation*}
$$

Proof: This follows from the definition of Lie derivative. If we have a metric $g$, a vector field $Y$ and $\lambda \in R$ such that

$$
-2 \operatorname{Ric}^{*}\left(g_{0}\right)=£_{Y} g_{0}-2 \lambda g_{0}
$$

after setting $g(t)=g_{0}$ and $\sigma(t)=1-2 \lambda t$ and then integrating the $t$-dependent vector filed $X(t)=\frac{1}{\sigma(t)} Y$. To give a family of deffeomorphism $\psi_{t}$ with $\psi_{0}$ the identity then $\bar{g}$ defined previously is a Ricci flow with

$$
\begin{gathered}
\bar{g}=g_{0} \frac{\partial \bar{g}}{\partial t}=\sigma^{\prime}(t) \phi_{t}^{*}\left(g_{0}\right)+\sigma(t) \phi_{t}^{*}\left(£_{X} g_{0}\right) \\
=\phi_{t}^{*}\left(-2 \lambda g_{0}+£_{Y} g_{0}\right)=\phi_{t}^{*}\left(-2 \operatorname{Ric}^{*}\left(g_{0}\right)\right)=-2 \operatorname{Ric}^{*}(\bar{g}) .
\end{gathered}
$$

Proposition 1.2. Under $*$-Ricci flow

$$
g\left(\frac{\partial}{\partial g} \nabla_{X} Y, Z\right)=-2\left(\nabla_{X} S^{*}\right)(Y, Z)+2 S^{*}\left(Y, \nabla_{X} Z\right)+2 S^{*}\left(\nabla_{X} Y, Z\right)
$$

Proof. Let us consider

$$
\frac{\partial}{\partial t} \nabla_{X} Y=\pi(X, Y)
$$

Now we can write

$$
\begin{equation*}
g\left(\frac{\partial}{\partial t} \nabla_{X} Y, Z\right)=g(\pi(X, Y), Z) \tag{1.5}
\end{equation*}
$$

Again

$$
\begin{align*}
g\left(\frac{\partial}{\partial t} \nabla_{X} Y, Z\right) & =\frac{\partial}{\partial t} g\left(\nabla_{X} Y, Z\right)-\frac{\partial g}{\partial t}\left(\nabla_{X} Y, Z\right) \\
g(\pi(X, Y), Z) & =\frac{\partial}{\partial t} g\left(\nabla_{X}, Z\right)+2 S^{*}\left(\nabla_{X} Y, Z\right) \tag{1.6}
\end{align*}
$$

We have

$$
\begin{equation*}
X g(Y, Z)=g\left(\nabla_{X} Y, Z\right)+g\left(Y, \nabla_{X} Z\right) \tag{1.7}
\end{equation*}
$$

From (1.5) we have

$$
\begin{aligned}
g(\pi(X, Y), Z) & =\frac{\partial}{\partial t}\left[X g(Y, Z)-g\left(Y, \nabla_{X} Z\right)\right]+2 S^{*}\left(\nabla_{X} Y, Z\right) \\
g(\pi(X, Y), Z) & =X \frac{\partial g}{\partial t}(Y, Z)-\left(\frac{\partial g}{\partial t}\right)\left(Y, \nabla_{X} Z\right)+2 S^{*}\left(\nabla_{X} Y, Z\right)
\end{aligned}
$$

or

$$
g\left(\pi(X, Y), Z=-2\left(\nabla_{X} S^{*}\right)(Y, Z)+2 S^{*}\left(Y, \nabla_{X} Z\right)+2 S^{*}\left(\nabla_{X} Y, Z\right)\right.
$$

i.e.

$$
\begin{equation*}
g\left(\frac{\partial}{\partial g} \nabla_{X} Y, Z\right)=-2\left(\nabla_{X} S^{*}\right)(Y, Z)+2 S^{*}\left(Y, \nabla_{X} Z\right)+2 S^{*}\left(\nabla_{X} Y, Z\right) \tag{1.8}
\end{equation*}
$$

## 2. The $\mathfrak{F}$-functional for the $*$-Ricci flow

Let $M$ be a fixed closed manifold, $g$ a Riemannian metric and $f$ a function defined on $M$ to the set of real numbers $\mathbb{R}$.

Then the $\mathfrak{F}$-functional on pair $(g, f)$ is defined as

$$
\begin{equation*}
\mathfrak{F}(g, f)=\int\left(-1+|\nabla f|^{2}\right) e^{-f} d V \tag{2.1}
\end{equation*}
$$

Now, we will establish how the $\mathfrak{F}$-functional changes according to time under $*$-Ricci flow.

Theorem 2.1. In $*$-Ricci flow the rate of change of $\mathfrak{F}$-functional with respect of time is given by

$$
\begin{aligned}
\frac{d}{d t} \mathfrak{F}(g, f)= & \int\left[-2 \operatorname{Ric}^{*}(\nabla f, \nabla f)-2 \frac{\partial f}{\partial t}\left(\Delta f-|\nabla f|^{2}\right)\right. \\
& \left.+\left(-1+|\nabla f|^{2}\right)\left(-\frac{\partial f}{\partial t}+\frac{1}{2} \operatorname{tr} \frac{\partial g}{\partial t}\right)\right] e^{-f} d V
\end{aligned}
$$

where

$$
\mathfrak{F}(g, f)=\int\left(-1+|\nabla f|^{2}\right) e^{-f} d V
$$

Proof. We may calculate

$$
\begin{equation*}
\frac{\partial}{\partial t}|\nabla f|^{2}=\frac{\partial}{\partial t} g(\nabla f, \nabla f)=\frac{\partial g}{\partial t}(\nabla f, \nabla f)+2 g\left(\nabla \frac{\partial f}{\partial t}, \nabla f\right) \tag{2.2}
\end{equation*}
$$

So using proposition 2.3.12 of [13] we can write

$$
\begin{align*}
\frac{d}{d t} \mathfrak{F}(g, f) & =\int\left[\frac{\partial g}{\partial t}(\nabla f, \nabla f)+2 g\left(\nabla \frac{\partial f}{\partial t}, \nabla f\right)\right] e^{-f} d V \\
& +\int\left(-1+|\nabla f|^{2}\right)\left[-\frac{\partial f}{\partial t}+\frac{1}{2} \operatorname{tr} \frac{\partial g}{\partial t}\right] e^{-f} d V \tag{2.3}
\end{align*}
$$

Using integration by parts of equation(2.2), we get

$$
\begin{equation*}
\int 2 g\left(\nabla \frac{\partial f}{\partial t}, \nabla f\right) e^{-f} d V=-2 \int \frac{\partial f}{\partial t}\left(\Delta f-|\nabla f|^{2}\right) e^{-f} d V \tag{2.4}
\end{equation*}
$$

Now putting (2.4) in (2.3), we get

$$
\begin{align*}
\frac{d}{d t} \mathfrak{F}(g, f)= & \int\left[\frac{\partial g}{\partial t}(\nabla f, \nabla f)-2 \frac{\partial f}{\partial t}\left(\Delta f-|\nabla f|^{2}\right)\right. \\
& \left.+\left(-1+|\nabla f|^{2}\right)\left(-\frac{\partial f}{\partial t}+\frac{1}{2} \operatorname{tr} \frac{\partial g}{\partial t}\right)\right] e^{-f} d V \tag{2.5}
\end{align*}
$$

Using (1.3) in (2.5), we get the following result for conformal Ricci flow, as

$$
\begin{align*}
\frac{d}{d t} \mathfrak{F}(g, f)=\int[ & -2 \operatorname{Ric}^{*}(\nabla f, \nabla f)-2 \frac{\partial f}{\partial t}\left(\Delta f-|\nabla f|^{2}\right) \\
& \left.+\left(-1+|\nabla f|^{2}\right)\left(-\frac{\partial f}{\partial t}+\frac{1}{2} \operatorname{tr} \frac{\partial g}{\partial t}\right)\right] e^{-f} d V \tag{2.6}
\end{align*}
$$

Hence the proof.

## 3. $\omega$-entropy functional for the $*$ - Ricci flow

Let $M$ be a closed manifold, $g$ a Riemannian metric on $M$ and $f$ a smooth function defined from $M$ to the set of real numbers $\mathbb{R}$. We define $\omega$-entropy functional as

$$
\begin{equation*}
\omega(g, f, \tau)=\int\left[\tau\left(R^{*}+|\nabla f|^{2}\right)+f-n\right] u d V \tag{3.1}
\end{equation*}
$$

where $\tau>0$ is a scale parameter and $u$ is defined as $u(t)=e^{-f(t)} ; \int_{M} u d V=1$.
We would also like to define heat operator acting on the function $f: M \times[0, \tau] \longrightarrow \mathbb{R}$ by $\diamond:=\frac{\partial}{\partial t}-\Delta$ and also, $\diamond^{*}:=-\frac{\partial}{\partial t}-\Delta+R^{*}$, conjugate to $\diamond$.

We choose $u$, such that $\diamond^{*} u=0$.
Now we prove the following theorem.

Theorem 3.1: If $g, f, \tau$ evolve according to

$$
\begin{equation*}
\frac{\partial g}{\partial t}=-2 R i c^{*} \tag{3.2}
\end{equation*}
$$

$$
\begin{gather*}
\frac{\partial \tau}{\partial t}=-1  \tag{3.3}\\
\frac{\partial f}{\partial t}=-\Delta f+|\nabla f|^{2}-R^{*}+\frac{n}{2 \tau} \tag{3.4}
\end{gather*}
$$

and the function $v$ is defined as $v=\left[\tau\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)+f-n\right] u$, the rate of change of $\omega$-entropy functional for conformal Ricci flow is $\frac{d \omega}{d t}=-\int_{M} \diamond^{*} v$, where

$$
\begin{aligned}
\diamond^{*} v= & 2 u\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{u n}{2 \tau}-v-u \tau\left[4<\text { Ric }^{*}, \text { Hess } f>\right. \\
& \left.-2 g\left(\nabla|\nabla f|^{2}, \nabla f\right)+4 g(\nabla(\Delta f), \nabla f)+2 \mid \text { Hess }\left.\right|^{2}\right]
\end{aligned}
$$

Proof: We find that

$$
\diamond^{*} v=\diamond^{*}\left(\frac{v}{u} u\right)=\frac{v}{u} \diamond^{*} u+u \diamond^{*}\left(\frac{v}{u}\right)
$$

We have defined previously that $\diamond^{*} u=0$,
so

$$
\begin{gathered}
\diamond^{*} v=u \diamond^{*}\left(\frac{v}{u}\right) \\
\diamond^{*} v=u \diamond^{*}\left[\tau\left(2 \nabla f-|\nabla f|^{2}+R^{*}\right)+f-n\right]
\end{gathered}
$$

We shall use the conjugate of heat operator, as defined earlier as $\diamond^{*}=-\left(\frac{\partial}{\partial t}+\Delta-R^{*}\right)$.

Therefore

$$
\begin{aligned}
& \diamond^{*} v=-u\left(\frac{\partial}{\partial t}\right.\left.+\Delta-R^{*}\right)\left[\tau\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)+f-n\right] \\
& \begin{aligned}
\Rightarrow u^{-1} \diamond^{*} v= & -\left(\frac{\partial}{\partial t}+\Delta\right)\left[\tau\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)\right] \\
& -\left(\frac{\partial}{\partial t}+\Delta\right) f-\left[\tau\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)+f-n\right] .
\end{aligned}
\end{aligned}
$$

Using equation (3.3), we have

$$
\begin{align*}
u^{-1} \diamond^{*} v=\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right) & -\tau\left(\frac{\partial}{\partial t}+\Delta\right)\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right) \\
& -\frac{\partial f}{\partial t}-\Delta f-\frac{v}{u} \tag{3.5}
\end{align*}
$$

Now

$$
\frac{\partial}{\partial t}\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)=2 \frac{\partial}{\partial t}(\Delta f)-\frac{\partial}{\partial t}|\nabla f|^{2}
$$

Using proposition (2.5.6) of [13], we have

$$
\begin{aligned}
\frac{\partial}{\partial t}\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)=2 \Delta & \frac{\partial f}{\partial t}
\end{aligned}+4<\text { Ric }^{*}, \text { Hess } f \gg 5\left(\frac{\partial g}{\partial t} \nabla f, \nabla f\right) .
$$

Now using the $*-$ Ricci flow equation (1.3), we have

$$
\begin{align*}
\frac{\partial}{\partial t}\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right) & =2 \Delta \frac{\partial f}{\partial t}+4<\text { Ric }^{*}, \text { Hess } f> \\
& +2 \text { Ric }^{*}(\nabla f, \nabla f)-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right) \tag{3.6}
\end{align*}
$$

Using (3.4) in (3.6), we get
$\frac{\partial}{\partial t}\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)=2 \Delta\left(-\Delta f+|\nabla f|^{2}-R^{*}+\frac{n}{2 \tau}\right)+4<$ Ric $^{*}$, Hessf $>$

$$
\begin{equation*}
+2 \operatorname{Ric}^{*}(\nabla f, \nabla f)-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right) \tag{3.7}
\end{equation*}
$$

Now let us compute

$$
\begin{equation*}
\Delta\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)=2 \Delta^{2} f-\Delta|\nabla f|^{2} \tag{3.8}
\end{equation*}
$$

Using (3.7) and (3.8) in (3.5) we obtain after a brief calculation

$$
\begin{aligned}
u^{-1} \diamond^{*} v & =\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)-\tau\left[-2 \Delta^{2} f+2 \Delta|\nabla f|^{2}+4<\text { Ric }^{*}, \text { Hess }>\right. \\
& \left.+2 \text { Ric }^{*}(\nabla f, \nabla f)-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right)+2 \Delta^{2} f-\Delta|\nabla f|^{2}\right]-\frac{\partial f}{\partial t}-\Delta f-\frac{v}{u} \\
& =\Delta f-|\nabla f|^{2}+R^{*}-\tau\left[\Delta|\nabla f|^{2}+4<\text { Ric }^{*}, \text { Hessf }>+2 \text { Ric }^{*}(\nabla f, \nabla f)\right. \\
& \left.-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right)\right]-\frac{\partial f}{\partial t}-\frac{v}{u} \\
& =\Delta f-|\nabla f|^{2}+R^{*}-\tau\left[\Delta|\nabla f|^{2}+4<\text { Ric }^{*}, \text { Hessf }>+2 \text { Ric }^{*}(\nabla f, \nabla f)\right. \\
& \left.-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right)\right]+\Delta f-|\nabla f|^{2}+R^{*}-\frac{n}{2 \tau}-\frac{v}{u} \\
& =2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-\frac{v}{u}-\tau\left[\Delta|\nabla f|^{2}+4<\text { Ric }^{*}, \text { Hessf }>\right. \\
& \left.+2 \text { Ric }^{*}(\nabla f, \nabla f)-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right)\right]
\end{aligned}
$$

$$
\begin{gathered}
u^{-1} \diamond^{*} v=2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-\left[\tau\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)+f-n\right]-\tau\left[\Delta|\nabla f|^{2}\right. \\
\left.+4<\text { Ric*}, \text { Hessf }>+2 \text { Ric*}(\nabla f, \nabla f)-2 g\left(\frac{\partial}{\partial t} \nabla f, \nabla f\right)\right] . \\
u^{-1} \diamond^{*} v=2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-f+n-\tau\left[2 \Delta f-|\nabla f|^{2}+R^{*}\right.
\end{gathered}
$$

$$
\begin{equation*}
\left.+\Delta|\nabla f|^{2}+4<\text { Ric }^{*}, \text { Hessf }>+2 \operatorname{Ric}^{*}(\nabla f, \nabla f)-2 g\left(\nabla \frac{\partial f}{\partial t}, \nabla f\right)\right] . \tag{3.9}
\end{equation*}
$$

Using (3.4), we get

$$
\begin{align*}
u^{-1} \diamond^{*} v= & 2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-f+n-\tau\left[2 \Delta f-|\nabla f|^{2}\right. \\
+ & R^{*}+\Delta|\nabla f|^{2}+4<\text { Ric }^{*}, \text { Hess } f>+2 \text { Ric }^{*}(\nabla f, \nabla f) \\
& \left.-2 g\left(\nabla\left(-\Delta f+|\nabla f|^{2}+\frac{n}{2 \tau}-R^{*}\right), \nabla f\right)\right] . \tag{3.10}
\end{align*}
$$

We can rewrite (3.10) in the following way

$$
\begin{align*}
u^{-1} \diamond^{*} v= & 2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-f+n-\tau\left[2 \Delta f-|\nabla f|^{2}+R^{*}\right. \\
& \left.+4<\text { Ric }^{*}, \text { Hessf }>-2 g\left(\nabla|\nabla f|^{2}, \nabla f\right)+4 g(\nabla(\Delta f), \nabla f)\right] \\
& +\tau\left[-\Delta|\nabla f|^{2}-2 \text { Ric }^{*}(\nabla f, \nabla f)+2 g(\nabla(\Delta f), \nabla f)\right] \tag{3.11}
\end{align*}
$$

and using Bochner formula in (3.11) and simplifying it, we get

$$
\begin{gathered}
u^{-1} \diamond^{*} v=2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-f+n-\tau\left[2 \Delta f-|\nabla f|^{2}+R^{*}\right. \\
+4<R i c^{*}, \text { Hess } f>-2 g\left(\nabla|\nabla f|^{2}, \nabla f\right) \\
+4 g(\nabla(\Delta f), \nabla f)]-2 \tau \mid \text { Hess }\left.f\right|^{2} . \\
\Rightarrow u^{-1} \diamond \diamond^{*} v=2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-\left[\tau\left(2 \Delta f-|\nabla f|^{2}+R^{*}\right)+f-n\right] \\
- \\
-\tau\left[4<\text { Ric } c^{*}, \text { Hess } f>-2 g\left(\nabla|\nabla f|^{2}, \nabla f\right)\right. \\
+4 g(\nabla(\Delta f), \nabla f)]-2 \tau \mid \text { Hessf }\left.\right|^{2} .
\end{gathered}
$$

i.e.

$$
\begin{align*}
u^{-1} \diamond^{*} v= & 2\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{n}{2 \tau}-\frac{v}{u}-\tau\left[4<\text { Ric }^{*}, \text { Hess } f>\right. \\
& \left.-2 g\left(\nabla|\nabla f|^{2}, \nabla f\right)+4 g(\nabla(\Delta f), \nabla f)\right]-2 \tau \mid \text { Hess }\left.f\right|^{2} . \tag{3.12}
\end{align*}
$$

So finally we have

$$
\begin{array}{r}
\diamond^{*} v=2 u\left(\Delta f-|\nabla f|^{2}+R^{*}\right)-\frac{u n}{2 \tau}-v-u \tau\left[4<\text { Ric }^{*}, \text { Hess } f>\right. \\
\left.-2 g\left(\nabla|\nabla f|^{2}, \nabla f\right)+4 g(\nabla(\Delta f), \nabla f)+2 \mid \text { Hessf }\left.\right|^{2}\right] . \tag{3.13}
\end{array}
$$

Now using remark (8.2.7) of [13], we get

$$
\frac{d \omega}{d t}=-\int_{M} \diamond^{*} v
$$

So the evolution of $\omega$ with respect to time can be found by this integration.
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Dipankar Debnath<br>Department of Mathematics<br>Bamanpukur High School(H.S)<br>Bamanpukur, Sree Mayapur<br>Nabadwip, Nadia<br>West Bengal, Pin-741313<br>India<br>dipankardebnath123@gmail.com

Nirabhra Basu
Department of Mathematics
Bhawanipur Education Society College
Kolkata-700020
West Bengal
India
nirabhra.basu@thebges.edu.in

# SURFACE FAMILY WITH COMMON LINE OF CURVATURE IN 3-DIMENSIONAL GALILEAN SPACE 

Mustafa Altin and İnan Ünal

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Abstract. In this paper we tried to find parametric presentation of a surface family with common line of curvature in 3 -dimensional Galilean space. We have obtained necessary and sufficient conditions for the curve to be a common line of curvature on this surface. We have stated examples to visualize our results and also, we have examined a torsion free curve.
Keywords: surface family; curvature; 3-dimensional Galilean space.

## 1. Introduction

The surface family (or pencil surface) is a notion in differential geometry applied in engineering science such as computer, manufacturing, mechanical engineering [25]. In 2004 Wang et al. [25] gave the definition of a surface family. Their paper is a reverse engineering problem to find a spatial curve to characterize the surface and also the paper contains conditions for a curve to be a geodesic on this surface. Besides, their work could be seen as an example of industrial mathematics. Kasap et al. [10] generalized this study by assumption of more general marchingscale functions. In [13] Li et al studied the approximation minimal surface with geodesics by using Dirichlet function and they minimized the area of surface family by using Dirichlet approach. This method can be used for obtaining minimal cost of material while building surfaces. The surface family notion has been studied by many researchers $[1,2,9,10]$.

There are many special curves on a surface such as geodesics. One of them is the line of curvature. A line of curvature is a curve on a surface whose tangent line at every point is aligned along a principal curvature direction. In [4] Che at al. analysed and computed these curves which are defined on implicit surface and worked on differential geometry of them. Same authors derived a necessary and

[^2]sufficient condition for a given curve to be the line of curvature on the surface. Surface family with common line of curvature has been studied in $[7,8,12]$.

Galileo geometry is a type of non-Euclidean geometry based on Galileo principle of relativity [20] and it has many important applications in physics [14]. In the last decades, these kind of spaces have become interesting by geometers because of their significant properties as a non-Euclidean geometry. Curves and surfaces in Galilean geometry has been studied by many authors $[3,5,6,15-17,21]$. Surfaces family, especially, in Galilean space have been studied in [22-24].

In this study, we examined a surface family with common line of curvature in 3 - dimensional Galilean space. We obtain necessary and sufficient conditions for the curve to be a line of curvature on the surface. We get some results for a torsion free curve. Finally, we present examples and plot their graphs.

## 2. Preliminaries

A. Cayley and F. Klein discovered that both Euclidean and non-Euclidean geometries can be considered as mathematical structures living inside projectivemetric spaces. Their contribution to geometry is called Cayley-Klein geometry and non- Euclidean geometries could be classified by this geometry. In fact, the 3-dimensional Galilean geometry is also a Cayley-Klein space [20].

### 2.1. Basic Facts in 3D Galilean Space

In this subsection, we recall some fundamental facts from Galilean geometry. For details see [18, 20].

A vector $\omega=\left(\omega_{1}, \omega_{2}, \omega_{3}\right)$ in 3-dimensional Galilean space $\mathbb{G}_{3}$ is called nonisotropic if $\omega_{1} \neq 0$, otherwise it is called isotropic.

Let $\omega=\left(\omega_{1}, \omega_{2}, \omega_{3}\right)$ and $\eta=\left(\eta_{1}, \eta_{2}, \eta_{3}\right)$ be two vectors in Galilean space $\mathbb{G}_{3}$. The inner product and the vector product of $\omega$ and $\eta$ in $\mathbb{G}_{3}$ are defined by

$$
\langle\omega, \eta\rangle=\left\{\begin{array}{c}
\omega_{1} \eta_{1}, \text { if } \omega_{1} \neq 0 \text { or } \eta_{1} \neq 0 \\
\omega_{2} \eta_{2}+\omega_{3} \eta_{3} \text { if } \omega_{1}=0 \text { and } \eta_{1}=0
\end{array}\right.
$$

and

$$
\omega \times \eta= \begin{cases}\left|\begin{array}{ccc}
e_{1} & e_{2} & e_{3} \\
0 & \omega_{2} & \omega_{3} \\
0 & \eta_{2} & \eta_{3}
\end{array}\right| & \text { if } \omega_{1}=\eta_{1}=0 \\
\left|\begin{array}{ccc}
0 & e_{2} & e_{3} \\
\omega_{1} & \omega_{2} & \omega_{3} \\
\eta_{1} & \eta_{2} & \eta_{3}
\end{array}\right| \quad \text { if } \omega_{1} \neq 0 \text { or } \eta_{1} \neq 0 .\end{cases}
$$

respectively.

Let $\gamma: I \rightarrow \mathbb{G}_{3}, \quad I \subset \mathbb{R}$ be a curve in $\mathbb{G}_{3}$ given by $\gamma(\phi)=(\delta(\phi), \zeta(\phi), \psi(\phi))$. Then the curvature $\kappa_{1}$ and torsion $\kappa_{2}$ of $\gamma(\phi)$ is obtained as
$(2.1) \kappa_{1}(\phi)=\left\|\gamma^{\prime \prime}(\phi)\right\|, \quad \kappa_{2}(\phi)=\frac{1}{\kappa_{1}^{2}(\phi)} \operatorname{det}\left(\gamma^{\prime}(\phi), \gamma^{\prime \prime}(\phi), \gamma^{\prime \prime \prime}(\phi)\right), \kappa_{1}(\phi) \neq 0$
where $\|$,$\| is the Galilean norm. Thus, we have Frenet formulas of \gamma(\phi)$ by

$$
\left\{\begin{array}{c}
V_{1}^{\prime}=\kappa_{1} V_{2},  \tag{2.2}\\
V_{2}^{\prime}=\kappa_{2} V_{3} \\
V_{3}^{\prime}=-\kappa_{2} V_{2}
\end{array}\right.
$$

where $V_{1}, V_{2}$ and $V_{3}$ are tangent, normal and binormal vector fields of $\gamma(\phi)$, respectively.
If $\delta^{\prime}(\phi)=0$, then $\gamma(\phi)$ is called a non-admissible curve, otherwise it is called an admissible curve. Let $\gamma(\phi)$ be an admissible curve in $\mathbb{G}_{3}$, given by

$$
\begin{equation*}
\gamma(\phi)=(\phi, \zeta(\phi), \psi(\phi)) \tag{2.3}
\end{equation*}
$$

Then $\kappa_{1}$ and $\kappa_{2}$ can be obtained as

$$
\kappa_{1}(\phi)=\sqrt{\zeta^{\prime \prime}(\phi)^{2}+\psi^{\prime \prime}(\phi)^{2}}, \kappa_{2}(\phi)=\frac{1}{\left(\kappa_{1}(\phi)\right)^{2}} \operatorname{det}\left(\gamma^{\prime}(\phi), \gamma^{\prime \prime}(\phi), \gamma^{\prime \prime \prime}(\phi)\right)
$$

and the Frenet vectors are given by

$$
\left\{\begin{array}{c}
V_{1}(\phi)=\gamma^{\prime}(\phi)=\left(1, \zeta^{\prime}(\phi), \psi^{\prime}(\phi)\right), \\
V_{2}(\phi)=\frac{\gamma^{\prime \prime}(\phi)}{\kappa_{1}(\phi)}=\frac{1}{\kappa_{1}(\phi)}\left(0, \zeta^{\prime \prime}(\phi), \psi^{\prime \prime}(\phi)\right) \\
V_{3}(\phi)=\frac{1}{\kappa_{1}(\phi)}\left(0,-\psi^{\prime \prime}(\phi), \zeta^{\prime \prime}(\phi)\right)
\end{array}\right.
$$

### 2.2. Some facts on Surface Theory in 3D Galilean Space

A surface in $\mathbb{G}_{3}$ is a parametric mapping from a region $R$ in $\mathbb{R}^{2}$ to $\mathbb{G}_{3}$ such as

$$
\begin{equation*}
S: R \subset \mathbb{R}^{2} \rightarrow \mathbb{G}_{3}, \quad S(\phi, \varphi)=\left(S_{1}(\phi, \varphi), S_{2}(\phi, \varphi), S_{3}(\phi, \varphi)\right) \tag{2.4}
\end{equation*}
$$

where $S_{1}, S_{2}$ and $S_{3}$ are functions in $C^{1}\left(\mathbb{G}_{3}, \mathbb{R}\right)$. The normal vector field of $S$ is given by

$$
\begin{equation*}
\mathcal{N}(\phi, \varphi)=S_{\phi} \times S_{\varphi} \tag{2.5}
\end{equation*}
$$

where $S_{\phi}=\frac{\partial S}{\partial \phi}$ and $S_{\varphi}=\frac{\partial S}{\partial \varphi}$ are partial derivatives of $S$.
Every surface has its own intrinsic geometry which has been known since Gauss. So, curves on a surface have geometric properties independent from the ambient space. We have a classification for curves on a surface by following definition.

Definition 2.1. Let $\gamma(\phi)$ be a curve on a surface $S$ in 3-dimensional Galilean space $\mathbb{G}_{3}$. Then $\gamma(\phi)$ is

1. a line of curvature, if the tangent vector at any point is in the direction of the principal curvature.
2. a geodesic if the normal vector field $V_{2}(\phi)$ of the curve $\gamma(\phi)$ and the normal $\mathcal{N}\left(\phi, \varphi_{0}\right)$ are parallel.
3. an asymptotic if the the binormal $V_{3}(\phi)$ of $\gamma(\phi)$ and the normal $\mathcal{N}\left(\phi, \varphi_{0}\right)$ of the surface at any point on $\gamma(\phi)$, are parallel to each other.

On the other hand, if $\gamma(\phi)$ is both an asymptotic and a parametric (isoparametric) curve, then it is called isoasymptotic; if it is both an geodesic and a parametric (isoparametric) curve, then it is called isogeodesic.

The well-known theorem below gives the conditions for any curve on a surface $S$ to be the line of curvature. For proof and details, we refer to reader [19].

Theorem 2.1. (Monge's Theorem) A necessary and sufficient condition for a curve on a surface to be a line of curvature is that the surface normals along the curve form a developable surface [19].

Let $S(\phi, \varphi)$ be a parametric surface in $\mathbb{G}_{3}$ is defined as follow;

$$
\begin{equation*}
S(\phi, \varphi)=\gamma(\phi)+\left[\lambda_{1}(\phi, \varphi) V_{1}(\phi)+\lambda_{2}(\phi, \varphi) V_{2}(\phi)+\lambda_{3}(\phi, \varphi) V_{3}(\phi)\right] \tag{2.6}
\end{equation*}
$$

for $(\phi, \varphi) \in R=\left[I_{1}, I_{2}\right] \times\left[I_{3}, I_{4}\right]$, where $\lambda_{1}(\phi, \varphi), \lambda_{2}(\phi, \varphi)$ and $\lambda_{3}(\phi, \varphi)$ are the values of the marching-scale functions in $C^{1}(S, \mathbb{R})$ and $\left\{V_{1}(\phi), V_{2}(\phi), V_{3}(\phi)\right\}$ is the Frenet frame of $\gamma(\phi)$. The surface (2.6) is called surface family with a common curve $\gamma(\phi)$.

A ruled surface formed by the surface normals can be given by

$$
\Psi(\phi, \varphi)=\gamma(\phi)+\varphi \mathfrak{n}
$$

where $\varphi$ is the distance of a point on $\Psi(\phi, \varphi)$ to point $\gamma(\phi)$ and $\mathfrak{n}=\cos \theta V_{2}(\phi)+$ $\sin \theta V_{3}(\phi)$, the vector functions $V_{2}(\phi), V_{3}(\phi)$ are the principal normal and the binormal of $\gamma(\phi)$, respectively. The surface $\Psi(\phi, \varphi)$ is called a normal surface [11].

Thus, by Monge's Theorem, $\gamma(\phi)$ is the line of curvature if and only if $\Psi(\phi, \varphi)$ is developable and $\mathfrak{n}$ is parallel to the normal vector field $\mathcal{N}$ of the surface (2.6). Also by classical differential geometry, it is well known that a surface is developable if and only if $\operatorname{det}\left(\gamma^{\prime}(\phi), \mathfrak{n}, \mathfrak{n}^{\prime}\right)=0$ (see [19]).

Hence from (2.2), we get

$$
\left|\begin{array}{ccc}
1 & 0 & 0  \tag{2.7}\\
0 & \cos \theta & \sin \theta \\
0 & -\theta^{\prime} \sin \theta-\kappa_{2} \sin \theta & \kappa_{2} \cos \theta+\theta^{\prime} \cos \theta
\end{array}\right|=0
$$

and so

$$
\theta^{\prime}+\kappa_{2}=0
$$

This means that

$$
\begin{equation*}
\theta=-\int_{\phi_{0}}^{\phi} \kappa_{2} d \phi+\theta_{0} \tag{2.8}
\end{equation*}
$$

where $\phi_{0}$ is the starting value of arc length and $\theta_{0}=\theta_{0}\left(\phi_{0}\right)$. In this paper, we assume $\phi_{0}=0$. Then by substituting $\theta$ in $\mathfrak{n}$ and with parallelity of $\mathfrak{n}$ to $\mathcal{N}$, we obtain the result that $\gamma(\phi)$ is the line of curvature.

## 3. Surfaces with common line of curvature in 3 D Galilean space $\mathbb{G}_{3}$

In this section, we work on surfaces family in Galilean 3 -space $\mathbb{G}_{3}$. We give if and only if conditions for a unit speed non-isotropic curve, being a line of curvature on a surface family. Furthermore, we give some examples and we present their graphics.

Theorem 3.1. The curve $\gamma(\phi)=(\phi, \zeta(\phi), \psi(\phi))$ is a line of curvature on the surface defined in (2.6) if and only if

$$
\left\{\begin{array}{c}
\lambda_{1}\left(\phi, \varphi_{0}\right)=\lambda_{2}\left(\phi, \varphi_{0}\right)=\lambda_{3}\left(\phi, \varphi_{0}\right)=0  \tag{3.1}\\
-\frac{\partial \lambda_{3}\left(\phi, \varphi_{0}\right)}{\partial \varphi}=\mu(\phi) \cos \theta, \frac{\partial \lambda_{2}\left(\phi, \varphi_{0}\right)}{\partial \varphi}=\mu(\phi) \sin \theta
\end{array}\right.
$$

where $(\phi, \varphi) \in R=\left[I_{1}, I_{2}\right] \times\left[I_{3}, I_{4}\right], \quad \mu(\phi) \neq 0$. The functions $\theta(\phi)$ and $\mu(\phi)$ are called controlling functions.

Proof. Let $S(\phi, \varphi)$ be a surface in $\mathbb{G}_{3}$ given by (2.6). For a curve $\gamma(\phi)$ on $S(\phi, \varphi)$ which is isoparametric, we have a parameter $\varphi_{0} \in\left[I_{3}, I_{4}\right]$ such that $\gamma(\phi)=S\left(\phi, \varphi_{0}\right)$ with conditions

$$
\lambda_{1}\left(\phi, \varphi_{0}\right)=\lambda_{2}\left(\phi, \varphi_{0}\right)=\lambda_{3}\left(\phi, \varphi_{0}\right)=0,\left(\phi, \varphi_{0}\right) \in R
$$

By direct computations, we have

$$
\begin{aligned}
\frac{\partial S(\phi, \varphi)}{\partial \phi} & =\left[1+\frac{\partial \lambda_{1}(\phi, \varphi)}{\partial \phi}\right] V_{1}(\phi) \\
& +\left[\kappa_{1} \lambda_{1}(\phi, \varphi)+\frac{\partial \lambda_{2}(\phi, \varphi)}{\partial \phi}-\kappa_{2} \lambda_{3}(\phi, \varphi)\right] V_{2}(\phi) \\
& +\left[\kappa_{2} \lambda_{2}(\phi, \varphi)+\frac{\partial \lambda_{3}(\phi, \varphi)}{\partial \phi}\right] V_{3}(\phi)
\end{aligned}
$$

and

$$
\frac{\partial S(\phi, \varphi)}{\partial \varphi}=\frac{\partial \lambda_{1}(\phi, \varphi)}{\partial \varphi} V_{1}(\phi)+\frac{\partial \lambda_{2}(\phi, \varphi)}{\partial \varphi} V_{2}(\phi)+\frac{\partial \lambda_{3}(\phi, \varphi)}{\partial \varphi} V_{3}(\phi)
$$

Thus, we get the normal vector of surface by

$$
\begin{equation*}
\mathcal{N}(\phi, \varphi)=\frac{\partial S(\phi, \varphi)}{\partial \phi} \times \frac{\partial S(\phi, \varphi)}{\partial \varphi} \tag{3.2}
\end{equation*}
$$

So, for $\varphi_{0} \in\left[I_{3}, I_{4}\right]$, we have

$$
\mathcal{N}\left(\phi, \varphi_{0}\right)=\mathcal{N}_{1}\left(\phi, \varphi_{0}\right) V_{1}(\phi)+\mathcal{N}_{2}\left(\phi, \varphi_{0}\right) V_{2}(\phi)+\mathcal{N}_{3}\left(\phi, \varphi_{0}\right) V_{3}(\phi)
$$

where

$$
\begin{aligned}
\mathcal{N}_{1}\left(\phi, \varphi_{0}\right) & =0 \\
\mathcal{N}_{2}\left(\phi, \varphi_{0}\right) & =\frac{\partial \lambda_{1}\left(\phi, \varphi_{0}\right)}{\partial \varphi}\left(\kappa_{2} \lambda_{2}\left(\phi, \varphi_{0}\right)+\frac{\partial \lambda_{3}\left(\phi, \varphi_{0}\right)}{\partial \phi}\right)-\left(1+\frac{\partial \lambda_{1}\left(\phi, \varphi_{0}\right)}{\partial \phi}\right) \frac{\partial \lambda_{3}\left(\phi, \varphi_{0}\right)}{\partial \varphi} \\
\mathcal{N}_{3}\left(\phi, \varphi_{0}\right) & =\left(1+\frac{\partial \lambda_{1}\left(\phi, \varphi_{0}\right)}{\partial \phi}\right) \frac{\partial \lambda_{2}\left(\phi, \varphi_{0}\right)}{\partial \varphi}-\left(\kappa_{1} \lambda_{1}\left(\phi, \varphi_{0}\right)\right. \\
& \left.+\frac{\partial \lambda_{2}\left(\phi, \varphi_{0}\right)}{\partial \phi}-\kappa_{2} \lambda_{3}\left(\phi, \varphi_{0}\right)\right) \frac{\partial \lambda_{1}\left(\phi, \varphi_{0}\right)}{\partial \varphi}
\end{aligned}
$$

Suppose that $\gamma(\phi)$ is a line of curvature on $S(\phi, \varphi)$. Thus, for a function $\mu(\phi) \neq 0$ on $S(\phi, \varphi)$, necessary and sufficient condition to provide $\mathfrak{n}(\phi) \| \mathcal{N}\left(\phi, \varphi_{0}\right)$ is

$$
\begin{aligned}
& \mathcal{N}_{2}\left(\phi, \varphi_{0}\right)=\mu(\phi) \cos \theta \\
& \mathcal{N}_{3}\left(\phi, \varphi_{0}\right)=\mu(\phi) \sin \theta
\end{aligned}
$$

So the proof is completed.
Example 3.1. Let $\gamma(\phi)=(\phi, 2 \sin (\phi), 2 \cos (\phi))$ be an admissible curve in $\mathbb{G}_{3}$. Then, we get the first and the second curvatures of $\gamma(\phi)$ by $\kappa_{1}=2$ and $\kappa_{2}=-1$. Thus, the Frenet frame is obtained by
$V_{1}(\phi)=(1,2 \cos (\phi),-2 \sin (\phi)), V_{2}(\phi)=(0,-\sin (\phi),-\cos (\phi)), V_{3}(\phi)=(0, \cos (\phi),-\sin (\phi))$.
If we choose

$$
\lambda_{1}(\phi, \varphi)=\phi^{2} \varphi, \lambda_{2}(\phi, \varphi)=\phi \sin (\theta) \sin (\phi \varphi), \lambda_{3}(\phi, \varphi)=\cos (\theta)\left(\phi-\phi e^{\phi \varphi}\right),
$$

then we get surface family $S(\phi, \varphi)$ given by (2.6) with common curve $\gamma(\phi)$. Then, by taking $\mu(\phi)=\phi^{2}$ and $\varphi_{0}=0$, the conditions given in (3.1) are satisfied.

Suppose that the normal surface $\Psi(\phi, \varphi)$ of $S$ is developable in $\mathbb{G}_{3}$. From (2.8), we have $\theta=\phi$. Thus, we get

$$
\begin{array}{ll}
S_{1}(\phi, \varphi) & =\phi+\phi^{2} \varphi \\
S_{2}(\phi, \varphi) & =2 \sin (\phi)+2 \phi^{2} \varphi \cos (\phi)+\sin ^{2}(\phi) \phi \sin (\phi \varphi)+\cos ^{2}(\phi)\left(\phi-\phi e^{\phi \varphi}\right) \\
S_{3}(\phi, \varphi) & =2 \cos (\phi)-2 \phi^{2} \varphi \sin (\phi)+\phi \sin (\phi) \cos (\phi) \sin (\phi \varphi)-\sin (\phi) \cos (\phi)\left(\phi-\phi e^{\phi \varphi}\right)
\end{array}
$$

As seen, all $S_{i}(\phi, \varphi), i=1,2,3$ are in $C^{1}(S, \mathbb{R})$. Consequently, $\gamma(\phi)$ is a line of curvature on $S(\phi, \varphi)$ with positive curvature and negative torsion.

By taking $R=[0,2 \pi] \times[-3,1]$, we visualize the curve $\gamma(\phi)=S(\phi, 0)$ in Fig.3.1, the surface $S(\phi, \varphi)$ in Fig. 3.2 and curve on surface in Fig. 3.3.


Fig. 3.1: Image of $\gamma(\phi)$


Fig. 3.2: Image of $S(\phi, \varphi)$


Fig. 3.3: $\gamma(\phi)$ on the $S(\phi, \varphi)$

Let take into consideration the case of the marching-scale functions

$$
\begin{aligned}
\lambda_{1}(\phi, \varphi) & =\rho_{1}(\phi) \Lambda_{1}(\varphi), \\
\lambda_{2}(\phi, \varphi) & =\rho_{2}(\phi) \Lambda_{2}(\varphi), \\
\lambda_{3}(\phi, \varphi) & =\rho_{3}(\phi) \Lambda_{3}(\varphi)
\end{aligned}
$$

with under conditions $\lambda_{1}\left(\phi, \varphi_{0}\right)=\lambda_{2}\left(\phi, \varphi_{0}\right)=\lambda_{3}\left(\phi, \varphi_{0}\right)=0$ and $\left(\phi, \varphi_{0}\right) \in R=$ $\left[I_{1}, I_{2}\right] \times\left[I_{3}, I_{4}\right]$, where $\rho_{1}(\phi), \Lambda_{1}(\varphi), \rho_{2}(\phi), \Lambda_{2}(\varphi), \rho_{3}(\phi)$ and $\Lambda_{3}(\varphi)$ are functions in $C^{1}(S, \mathbb{R})$. Then from Theorem 3.1, we have following corollary:

Corollary 3.1. The curve $\gamma(\phi)=(\phi, \zeta(\phi), \psi(\phi))$ is a line of curvature on the surface defined in (2.6) if and only if

$$
\left\{\begin{array}{c}
\Lambda_{1}\left(\varphi_{0}\right)=\Lambda_{2}\left(\varphi_{0}\right)=\Lambda_{3}\left(\varphi_{0}\right)=0  \tag{3.3}\\
-\rho_{3}(\phi) \frac{d \Lambda_{3}}{d \varphi}\left(\varphi_{0}\right)=\mu(\phi) \cos \theta, \rho_{2}(\phi) \frac{d \Lambda_{2}}{d \varphi}\left(\varphi_{0}\right)=\mu(\phi) \sin \theta
\end{array}\right.
$$

where $\left(\phi, \varphi_{0}\right) \in R=\left[I_{1}, I_{2}\right] \times\left[I_{3}, I_{4}\right]$ and $\mu(\phi) \neq 0$.
Example 3.2. Let $\gamma(\phi)=(\phi, \cos (\phi), \sin (\phi))$ be an admissible curve in $\mathbb{G}_{3}$. Then, we get first two curvatures as $\kappa_{1}=1$ and $\kappa_{2}=1$. Also the Frenet frame is given by
$V_{1}(\phi)=(1,-\sin (\phi), \cos (\phi)), \quad V_{2}(\phi)=(0,-\cos (\phi),-\sin (\phi)), \quad V_{3}(\phi)=(0, \sin (\phi), \cos (\phi))$.
Thus, we get surface family $S(\phi, \varphi)$ given by (2.6) with common curve $\gamma(\phi)$. Suppose that the normal surface $\Psi(\phi, \varphi)$ of $S$ is developable in $\mathbb{G}_{3}$. Thus, from (2.8), we have $\theta=-\phi$.

If we choose $\rho_{1}(\phi)=\phi, \Lambda_{1}(\phi, \varphi)=\left(\varphi^{2}-1\right), \rho_{2}(\phi)=\rho_{3}(\phi)=1, \Lambda_{2}(\phi, \varphi)=\sin (\theta)(\varphi-$ 1), $\Lambda_{3}(\phi, \varphi)=\cos (\theta)(1-\varphi)$, and take $\mu(\phi)=1, \varphi_{0}=1$ so that equation (3.1) is satisfied, then a member of surface family in $\mathbb{G}_{3}$ is obtained by

$$
\begin{aligned}
S(\phi, \varphi)= & \left(\phi+\phi\left(\varphi^{2}-1\right), \cos (\phi)-\phi\left(\varphi^{2}-1\right) \sin (\phi),\right. \\
& \left.\sin (\phi)+\phi\left(\varphi^{2}-1\right) \cos (\phi)+\sin ^{2}(\phi)(\varphi-1)+\cos ^{2}(\phi)(1-\varphi)\right)
\end{aligned}
$$

By taking $R=[0,2 \pi] \times[0,3]$, we visualize the curve $\gamma(\phi)=S(\phi, 0)$ in Fig.3.4, the surface $S(\phi, \varphi)$ in Fig. 3.5 and curve on surface in Fig. 3.6.


Fig. 3.4: Image of $\gamma(\phi)$


Fig. $\quad 3.5$ : Image of $S(\phi, \varphi)$


Fig. 3.6: $\gamma(\phi)$ on $S(\phi, \varphi)$

Suppose that the second curvature of $\gamma(\phi)$ vanish, i.e $\kappa_{2}=0$. Then, from (2.8), we have $\theta=\theta_{r}$ (constant). Thus, from (3.3), we obtain

$$
\frac{\mu(\phi)}{\rho_{3}(\phi)}=-c_{1}, \quad \frac{\mu(\phi)}{\rho_{2}(\phi)}=c_{2}
$$

Considering conditions in (3.3), we get

$$
\frac{d \Lambda_{3}}{d \varphi}\left(\varphi_{0}\right)=c_{1} \cos \theta_{r} \text { and } \frac{d \Lambda_{2}}{d \varphi}\left(\varphi_{0}\right)=c_{2} \cos \theta_{r}
$$

On the other hand, since $\mathfrak{n} \| \mathcal{N}$, if $\theta_{r}=(2 m+1) \frac{\pi}{2}$ for any integer $m$ then $V_{3} \| \mathcal{N}$. Thus, $\gamma(\phi)$ is an isoasymptotic curve on the surface. Also, if $\theta_{r}=m \pi$ then $V_{2} \| \mathcal{N}$ meaning $\gamma(\phi)$ is an isogeodesic curve on the surface. Consequently, we obtain the following result.

Corollary 3.2. Let the curve $\gamma(\phi)=(\phi, \zeta(\phi), \psi(\phi))$ be a line of curvature with torsion free on the surface is defined in (2.6). Then, we have

$$
\begin{aligned}
& \text { if } \theta_{r}=(2 m+1) \frac{\pi}{2} \text { for any integer } m \text {, then } \gamma(\phi) \text { is also isoasympotic, } \\
& \text { if } \theta_{r}=m \pi \text { for any integer } m \text {, then } \gamma(\phi) \text { is also isogeodesic. }
\end{aligned}
$$

Example 3.3. Let $\gamma(\phi)=(\phi, 1+\sin \phi, \sin \phi)$ be an admissible curve in $\mathbb{G}_{3}$. Then, we get first two curvatures as $\kappa_{1}=\sqrt{2} \sin \phi$ and $\kappa_{2}=0$. Also, the Frenet frame is given by

$$
V_{1}(\phi)=(1, \cos (\phi), \cos (\phi)), \quad V_{2}(\phi)=\left(0,-\frac{1}{\sqrt{2}},-\frac{1}{\sqrt{2}}\right), \quad V_{3}(\phi)=\left(0, \frac{1}{\sqrt{2}},-\frac{1}{\sqrt{2}}\right) .
$$

If we choose

$$
\rho_{1}(\phi)=\phi^{2}, \Lambda_{1}(\varphi)=\varphi, \quad \rho_{2}(\phi)=\phi, \Lambda_{2}(\varphi)=\varphi \sin \left(\theta_{r}\right), \quad \rho_{3}(\phi)=\phi, \Lambda_{3}(\phi)=-\varphi \cos \theta_{r}
$$

and take $\varphi_{0}=0, c_{1}=c_{2}=1$, then a member of surface family in $\mathbb{G}_{3}$ is obtained by

$$
\begin{aligned}
S(\phi, \varphi)= & \left(\phi+\phi^{2} \varphi, 1+\sin \phi-\frac{1}{\sqrt{2}} \phi \varphi \sin \left(\theta_{r}\right)-\frac{1}{\sqrt{2}} \phi \varphi \cos \left(\theta_{r}\right)\right. \\
& \left.\sin \phi-\frac{1}{\sqrt{2}} \phi \varphi \sin \theta_{r}+\frac{1}{\sqrt{2}} \phi \varphi \cos \theta_{r}\right) .
\end{aligned}
$$

By taking $R=[0,2 \pi] \times[0,0.2]$, we visualize the curve $\gamma(\phi)=S(\phi, 0)$ in Fig.3.7 and

1. the surface $S(\phi, \varphi)$ in Fig. 3.8 for $\theta=\frac{\pi}{6}$;
2. curve on surface in Fig. 3.9 for $\theta=\frac{\pi}{6}$, in Fig. 3.10 for $\theta=\frac{\pi}{2}$; in Fig. 3.11 for $\theta=0$.


FIg. 3.7: Image of $\gamma(\phi)$


Fig. 3.8: Image of $S(\phi, \varphi)$


FIg. 3.10: $\gamma(\phi)$ on $S(\phi, \varphi)$

FIG. 3.9: $\gamma(\phi)$ on $S(\phi, \varphi)$


FIG. 3.11: $\gamma(\phi)$ on $S(\phi, \varphi)$

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Mustafa Altin
Technical Sciences Vocational School
Bingol University
12000 Bingöl, Turkey
maltin@bingol.edu.tr

İnan Ünal
Department of Computer Engineering
Munzur University
62000 Tunceli, Turkey
inanunal@munzur.edu.tr

# A CLASSIFICATION OF SOME ALMOST $\alpha$-PARA-KENMOTSU MANIFOLDS * 

Quanxiang Pan and Ximin Liu


#### Abstract

In this paper, we mainly study local structures and curvatures of the almost $\alpha$-para-Kenmotsu manifolds. In particular, locally symmetric almost $\alpha$-para-Kenmotsu manifolds satisfying certain nullity conditions are classified.


Key words: curvatures; $\alpha$-para-Kenmotsu manifolds; nullity conditions.

## 1. Introduction

One of the recent topics in the theory of almost contact metric manifolds is the study of the so-called nullity distributions. In [5], E. Boeckx studied the full classification of contact ( $\kappa, \mu$ )-spaces, later in [11] and [12], P. Dacko and Z. Olszak gave a systematic study of almost cosymplectic ( $\kappa, \mu, \nu$ )-spaces and almost cosymplectic $(-1, \mu, 0)$-spaces. G. Dileo and A. M. Pastore in [8] studied nullity distributions on almost Kenmotsu manifolds. In recent years, many authors have turned to the study of almost paracontact geometry due to an unexpected relationship between contact $(\kappa, \mu)$-spaces and paracontact geometry that was found in [3].

The study of almost paracontact geometry was introduced by Kaneyuki and Williams in [14] and then it was continued by many other authors. A systematic study of almost paracontact metric manifolds was carried out in [16] by Zamkovoy. In fact, such manifolds were studied earlier in [17],[18],[6],[15] and in these papers the authors called such structures almost para-cohermitian. The curvature identities for different classes of almost paracontact metric manifolds were obtained in [13],[10], [16].

In [2], a complete study of paracontact metric manifolds satisfying a certain nullity condition has been carried out, later, in [9], the authors gave a complete study of almost $\alpha$-cosymplectic manifolds, where $\alpha$ is a function, basic properties of such manifolds are obtained and general curvature identities are proved. It is

[^3]also showed that almost $\alpha$-para-Kenmotsu $(\kappa, \mu, \nu)$-spaces have para-Kähler leaves. Motivated by [7], [8] and [9], the aim of this paper is devoted to investigate local symmetry and nullity distributions on almost $\alpha$-para-Kenmotsu manifolds.

This paper is organized in the following way. In section 2, some preliminaries and properties about almost $\alpha$-para-Kenmotsu manifolds are given. In section 3, we characterize almost paracontact metric manifolds which are $\mathcal{C R}$-integrable almost $\alpha$-para-Kenmotsu through the existence of a suitable linear connection, and in section 4 , we investigate almost $\alpha$-para-Kenmotsu manifolds which are locally symmetric and give some properties. In section 5, we study almost $\alpha$-para-Kenmotsu manifolds satisfying some nullity distributions and give some properties and classification theorems of them.

## 2. Almost $\alpha$-para-Kenmotsu manifolds

Now, we recall some basic notions of almost paracontact manifold (see [9] ). A $2 n+1$ dimensional smooth manifold $M$ is said to have an almost paracontact structure if it admits a $(1,1)$-tensor field $\varphi$, a vector field $\xi$ and a 1-form $\eta$ satisfying the following conditions:
(i) $\varphi^{2}=\mathrm{Id}-\eta \otimes \xi, \quad \eta(\xi)=1$,
(ii) the tensor field $\varphi$ induces to an almost paracomplex structure on each fibre of $\mathcal{D}=\operatorname{Ker}(\eta)$, i.e. the $\pm 1$-eigendistributions $\mathcal{D}^{ \pm}:=\mathcal{D}_{\varphi}( \pm 1)$ of $\varphi$ have equal dimension $n$.

From the definition, it follows that $\varphi(\xi)=0, \eta \circ \varphi=0$ and $\operatorname{rank}(\varphi)=2 n$. When the tensor field $\mathcal{N}_{\varphi}:=[\varphi, \varphi]-2 d \eta \otimes \xi$ vanishes identically the almost paracontact manifold is said to be normal. If an almost paracontact manifold admits a pseudoRiemannian metric $g$ such that

$$
\begin{equation*}
g(\varphi X, \varphi Y)=-g(X, Y)+\eta(X) \eta(Y) \tag{2.1}
\end{equation*}
$$

for any vector fields $X, Y \in \Gamma(T M)$, then we say that $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ is an almost paracontact metric manifold. Notice that any such a pseudo-Riemannian metric is necessarily of signature $(n, n+1)$. For an almost paracontact metric manifold, there always exists an orthogonal basis $\left\{\xi, X_{1}, \ldots, X_{n}, Y_{1}, \ldots, Y_{n}\right\}$ such that $g\left(X_{i}, X_{j}\right)=\delta_{i j}, g\left(Y_{i}, Y_{j}\right)=-\delta_{i j}$ and $Y_{i}=\varphi X_{i}$, for any $i, j \in\{1, \ldots, n\}$. Such basis is called a $\varphi$-basis. Moreover, we can define a skew-symmetric tensor field 2-form $\Phi$ by $\Phi(X, Y):=g(X, \varphi Y)$, which is usually called the fundamental form.

Lemma 2.1. ([16]) For an almost paracontact structure $(\varphi, \xi, \eta, g)$, the covariant derivative $\nabla \varphi$ of $\varphi$ with respect to the Levi-Civita connection $\nabla$ is given by

$$
\begin{aligned}
2 g\left(\left(\nabla_{X} \varphi\right) Y, Z\right)= & -3 d \Phi(X, \varphi Y, \varphi Z)-3 d \Phi(X, Y, Z)-g\left(\mathcal{N}^{(1)}(Y, Z), \varphi X\right) \\
& +\mathcal{N}^{(2)}(Y, Z) \eta(X)+2 d \eta(\varphi Y, X) \eta(Z)-2 d \eta(\varphi Z, X) \eta(Y)
\end{aligned}
$$

Definition 2.1. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost paracontact metric manifold, if it satisfies

$$
d \eta=0, \quad d \Phi=2 \alpha \eta \wedge \Phi
$$

where $\alpha=$ const. $\neq 0$, then $M^{2 n+1}$ is called an almost $\alpha$-para-Kenmotsu manifold.
Let $M$ be an almost $\alpha$-para-Kenmotsu manifold with structure $(\varphi, \xi, \eta, g)$. Since the 1 -form $\eta$ is closed, then the distribution $\mathcal{D}=\operatorname{ker}(\eta)$ is integrable, we have $L_{\xi} \eta=0$, and $[X, \xi] \in \mathcal{D}$ for any $X \in \mathcal{D}$. Then, using Lemma 2.1, the Levi-Civita connection is given by

$$
\begin{equation*}
2 g\left(\left(\nabla_{X} \varphi\right) Y, Z\right)=-2 \alpha g(\eta(Y) \varphi X+g(X, \varphi Y) \xi, Z)-g(\mathcal{N}(Y, Z), \varphi X) \tag{2.2}
\end{equation*}
$$

for any vector fields $X, Y, Z \in \Gamma(T M)$. If we replace $X$ by $\xi$, it follows $\nabla_{\xi} \varphi=0$, which implies that $\nabla_{\xi} \xi=0$ and $\nabla_{\xi} X \in \mathcal{D}$ for any $X \in \mathcal{D}$.

The tensor fields $h=\frac{1}{2} \mathcal{L}_{\xi} \varphi$ and $h^{\prime}=h \cdot \varphi$ are symmetric operators anticommuting with $\varphi$ and $h \xi=0=h^{\prime} \xi$, and we note that $\nabla_{\xi} h^{\prime}=0$ if and only if $\nabla_{\xi} h=0$. Let $Y=\xi$ in (2.2) we obtain

$$
\begin{equation*}
\nabla_{X} \xi=\alpha \varphi^{2} X+\varphi h X \tag{2.3}
\end{equation*}
$$

Proposition 2.1. An almost $\alpha$-para-Kenmotsu manifold $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ has para-Kähler leaves if and only if

$$
\left(\nabla_{X} \varphi\right) Y=g(\alpha \varphi X+h X, Y)-\eta(Y)(\alpha \varphi X+h X)
$$

Theorem 2.1. ([9]) Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold with para-Kähler leaves. Then $M^{2 n+1}$ is para-Kenmotsu $(\alpha=1)$ if and only if $\nabla_{X} \xi=\varphi^{2} X$.

Proposition 2.2. ([9]) Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold. Then, for any $X, Y, Z \in \Gamma\left(T M^{2 n+1}\right)$,

$$
\begin{gather*}
R(\xi, X) \xi=\alpha^{2} \varphi^{2} X+2 \alpha \varphi h X-h^{2} X+\varphi\left(\nabla_{\xi} h\right) X,  \tag{2.4}\\
\frac{1}{2}(R(\xi, X) \xi+\varphi R(\xi, \varphi X) \xi)=\alpha^{2} \varphi^{2} X-h^{2} X,  \tag{2.5}\\
R(X, Y) \xi=\alpha \eta(X)(\alpha Y+\varphi h Y)  \tag{2.6}\\
-\alpha \eta(Y)(\alpha X+\varphi h X)+\left(\nabla_{X} \varphi h\right) Y-\left(\nabla_{Y} \varphi h\right) X, \\
g(R(\xi, X) Y, Z)+g(R(\xi, X) \varphi Y, \varphi Z)-g(R(\xi, \varphi X) \varphi Y, Z) \\
-g(R(\xi, \varphi X) Y, \varphi Z)=2\left(\nabla_{h X} \Phi\right)(Y, Z)+2 \alpha^{2} \eta(Y) g(X, Z)  \tag{2.7}\\
-2 \alpha^{2} \eta(Z) g(X, Y)-2 \alpha \eta(Z) g(\varphi h X, Y)+2 \alpha \eta(Y) g(\varphi h X, Z) .
\end{gather*}
$$

Proposition 2.3. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold. Then, for any $X, Y \in \Gamma\left(T M^{2 n+1}\right)$,

$$
\begin{equation*}
g(\mathcal{N}(\varphi X, Y), \xi)=0 \tag{2.8}
\end{equation*}
$$

Proof. By direct computations one has

$$
g(\mathcal{N}(\varphi X, Y), \xi)=g([\varphi X, \varphi Y], \xi)=g\left(\left(\nabla_{X} \varphi\right) Y-\left(\nabla_{\varphi Y} \varphi\right) \varphi X, \xi\right)
$$

which implies (2.8) by using (2.2) and $[\xi, X]=-2 \varphi h X$.
Theorem 2.2. ([9]) Let $M^{2 n+1}$ be an almost $\alpha$-para-Kenmotsu manifold with $h=$ 0 . Then, $M^{2 n+1}$ is locally a warped product $M_{1} \times{ }_{f^{2}} M_{2}$, where $M_{2}$ is an almost para-Kähler manifold, $M_{1}$ ia an open interval with coordinate $t$ and $f^{2}=w e^{2 \alpha t}$ for some positive constant.

## 3. $\mathcal{C} \mathcal{R}$-integrability

For an almost $\alpha$-para-Kenmotsu manifold we have $[X, Y]-[\varphi X, \varphi Y] \in \mathcal{D}$ for any $X, Y \in \mathcal{D}$, since $d \eta=0$ and thus $\mathcal{D}$ is integrable. Hence, the structure $(\varphi, \xi, \eta, g)$ is $\mathcal{C R}$-integrable if and only if $\mathcal{N}(X, Y)=[X, Y]+[\varphi X, \varphi Y]-\varphi[\varphi X, Y]-\varphi[X, \varphi Y]=0$ on $\mathcal{D}$, that is to the request that the integral manifolds of $\mathcal{D}$ are para-Kähler.

Theorem 3.1. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost paracontact metric manifold. Then, $M^{2 n+1}$ is a $\mathcal{C R}$-integrable almost $\alpha$-para-Kenmotsu manifold if and only if there exists a linear connection $\tilde{\nabla}$ such that

1) $\tilde{\nabla} \varphi=0, \tilde{\nabla} g=0, \tilde{\nabla} \eta=0$.
2) the torsion $\tilde{T}$ satisfies
a) $\tilde{T}(X, Y)=0$ for any $X, Y \in D$,
b) $\tilde{T}(\xi, X)=X+h^{\prime} X$ for any $X \in D$,
c) $\tilde{T}_{\xi}$ is selfadjoint.

Moreover, such a connection is uniquely determined by

$$
\begin{equation*}
\tilde{\nabla}_{X} Y=\nabla_{X} Y+g\left(\alpha X-h^{\prime} X, Y\right) \xi-\eta(Y)\left(\alpha X-h^{\prime} X\right) \tag{3.1}
\end{equation*}
$$

$\nabla$ being the Levi-Civita connection.
$\underset{\tilde{\nabla}}{\text { Proof. Let }} M^{2 n+1}$ is a $\mathcal{C} \mathcal{R}$-integrable almost $\alpha$-para-Kenmotsu manifold. We put $\tilde{\nabla}=\nabla+H$, where the tensor field $H$ of type $(1,2)$ is defined by

$$
H(X, Y)=g\left(\alpha X-h^{\prime} X, Y\right) \xi-\eta(Y)\left(\alpha X-h^{\prime} X\right)
$$

Since $H(X, \varphi Y)-\varphi(H(X, Y))=-(g(\alpha \varphi X+h X, Y)-\eta(Y)(\alpha \varphi X+h X))=$ $-\left(\nabla_{X} \varphi\right) Y$, owing to Proposition 2.1. By direct calculations, we get $g(H(X, Y), Z)+$ $g_{\tilde{\nabla}}(H(X, Z), Y)=0$ and $\left(\nabla_{X} \eta\right) Y-\eta(H(X, Y))=0$, moreover, we get $\tilde{\nabla} \varphi=0$, $\tilde{\nabla} g=0, \tilde{\nabla} \eta=0$. Since $\tilde{T}(X, Y)=\eta(X)\left(\alpha Y-h^{\prime} Y\right)-\eta(Y)\left(\alpha X-h^{\prime} X\right)=0$ for any $X, Y \in \mathcal{D}$, and $\tilde{T}(\xi, X)=\alpha X-h^{\prime} X$ for any $X \in \mathcal{D}$, hence $\tilde{T}_{\xi}$ is selfadjoint. As for the uniqueness and the vice versa part, the proof is similar with Theorem 3.1 in [8].

Corollary 3.1. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be a $\mathcal{C R}$-integrable almost $\alpha$-para-Kenmotsu manifold. Then $M^{2 n+1}$ is a $\alpha$-para-Kenmotsu manifold if and only if the linear connection $\tilde{\nabla}$ verifies $\tilde{T}_{\xi} \circ \varphi=\varphi \circ \tilde{T}_{\xi}$.

Proof. Since $\tilde{T}_{\xi} \varphi X-\varphi \tilde{T}_{\xi} X=\tilde{T}(\xi, \varphi X)-\varphi \tilde{T}(\xi, X)=-2 h X$ for any $X \in \mathcal{D}$, hence, Corollary 3.1 is easily followed by Theorem 3.1.

## 4. Almost $\alpha$-para-Kenmotsu manifolds and local symmetrys

In this section, we investigate almost $\alpha$-para-Kenmotsu manifolds which are locally symmetric, that is, almost $\alpha$-para-Kenmotsu manifolds satisfying the condition $\nabla R=0$, which is a natural generalization of almost $\alpha$-para-Kenmotsu manifold of constant curvature.

By similar proof as that of proposition 6 in [7], we get the following lemma
Lemma 4.1. Let $M^{2 n+1}$ be a locally symmetric almost $\alpha$-para-Kenmotsu manifold. Then, $\nabla_{\xi} h=0$.

Theorem 4.1. $\operatorname{Let}\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be a locally symmetric almost $\alpha$-para-Kenmotsu manifold. Then, $M^{2 n+1}$ is an $\alpha$-para-Kenmotsu manifold if and only if $h=0$. Moreover, if any of the above equivalent conditions holds, $M^{2 n+1}$ has constant sectional curvature $c=-\alpha^{2}$.

Proof. First, assuming that $M^{2 n+1}$ is an $\alpha$-para-Kenmotsu manifold, by Theorem 2.1. we have $\nabla_{X} \xi=\alpha \varphi^{2} X$, comparing with (2.3) it follows that $h=0$ and by (2.6), we easily obtain $R(X, Y) \xi=-\alpha^{2}(\eta(Y) X-\eta(X) Y)$, let $\nabla_{Z}$ acting on the above equation and by the local symmetry, we have $R(X, Y) Z=-\alpha^{2}(g(Y, Z) X-$ $g(X, Z) Y)$, it follows then $M$ is of constant sectional curvature $c=-\alpha^{2}$. Now, supposing $M^{\prime}$ is the integral manifold of $\mathcal{D}$ and $\nabla^{\prime}$ is the corresponding connection on $M^{\prime}$. Then $\nabla_{X} Y=\nabla_{X}^{\prime} Y+h(X, Y)$, then $h(X, Y)=g\left(\nabla_{X} Y, \xi\right) \xi=-\alpha g(X, Y) \xi$, this implies $H=-\alpha \xi$ thus $h(X, Y)=g(X, Y) H$, and $M^{\prime}$ is a totally umbilical submanifold of $M^{2 n+1}$. What is more, it is not difficult to see that $R^{\prime}(X, Y)=$ $R(X, Y)+\alpha^{2}(g(Y, Z) X-g(X, Z) Y)=0$, we know that $M^{\prime}$ is flat and the sectional curvature of $M^{\prime}$ vanishes. This means that $M^{\prime}$ is a flat para-Kähler manifold. For another part of the proof, noticing that $\nabla_{Z} \xi=\alpha \varphi^{2} Z=\alpha Z$ if and only if $h=0$, by Theorem 2.1 we prove that $M^{2 n+1}$ is an $\alpha$-para-Kenmotsu manifold. At last, it is obvious from the proof of the equivalence that if any of the above equivalent conditions holds, $M^{2 n+1}$ has constant sectional curvature $c=-\alpha^{2}$. Thus, we complete the proof.

Theorem 4.2. An almost $\alpha$-para-Kenmotsu manifold of constant curvature $c$ is an $\alpha$-para-Kenmotsu manifold and $c=-\alpha^{2}$.

Proof. Supposing $M^{2 n+1}$ is an almost $\alpha$-para-Kenmotsu manifold of constant sectional curvature $c$, it is obvious that

$$
\begin{equation*}
R(X, Y) Z=c(\eta(Y) X-\eta(X) Y) \tag{4.1}
\end{equation*}
$$

$\nabla_{W}$ acting on (4.1) we get $\nabla_{W} R=0$, thus, $M^{2 n+1}$ is locally symmetric, by Lemma 4.1, we get $\nabla_{\xi} h=0$. Comparing (2.6) with (4.1), we obtain
$\left(c+\alpha^{2}\right)(\eta(Y) X-\eta(X) Y)+\alpha(\eta(Y) \varphi h X-\eta(X) \varphi h Y)-\left(\nabla_{X} \varphi h\right) Y+\left(\nabla_{Y} \varphi h\right) X=0$.
Choosing $X=\xi$ and $Y \in \mathcal{D}$ and by Lemma 4.1, we get

$$
\begin{equation*}
-\left(c+\alpha^{2}\right) Y-2 \alpha \varphi h Y+h^{2} X=0 \tag{4.2}
\end{equation*}
$$

Now, if $Y$ is an eigenvector of $h$ with eigenvalue $\lambda$, then (4.2) becomes $-(c+$ $\left.\alpha^{2}\right) Y-2 \alpha \lambda \varphi Y+\lambda^{2} X=0$. We get $\lambda=0$ and $c=-\alpha^{2}$ since $Y$ and $\varphi Y$ are linearly independent. Hence $h=0$ and $c=-\alpha^{2}$, by Theorem 4.1, we know $M^{2 n+1}$ is an $\alpha$-para-Kenmotsu manifold of constant curvature $c=-\alpha^{2}$. Thus, we complete the proof.

## 5. Almost $\alpha$-para-Kenmotsu manifolds and nullity distributions

In this section, we study almost $\alpha$-para-Kenmotsu manifolds under the assumption that $\xi$ belongs to the $(\kappa, \mu)$-nullity distribution and $(\kappa, \mu)^{\prime}$-nullity distribution.

First, we consider the $(\kappa, \mu)$-nullity distribution. if $\xi$ belongs to the $(\kappa, \mu)$ nullity distribution, $(\kappa, \mu) \in R^{2}$, denoted by $\mathcal{N}(\kappa, \mu)$, which is given by putting for each $p \in M^{2 n+1}$,

$$
\begin{aligned}
\mathcal{N}_{p}(\kappa, \mu) & =\left\{Z \in \Gamma\left(T_{p} M^{2 n+1}\right) \mid R(X, Y) Z\right. \\
& =\kappa(g(Y, Z) X-g(X, Z) Y)+\mu(g(Y, Z) h X-g(X, Z) h Y)\}
\end{aligned}
$$

So, if $\xi \in \mathcal{N}(\kappa, \mu)$, that is, for any $\mathrm{X}, \mathrm{Y} \in \Gamma\left(T M^{2 n+1}\right)$

$$
R(X, Y) \xi=\kappa(\eta(Y) X-\eta(X) Y)+\mu(\eta(Y) h X-\eta(X) h Y)
$$

Proposition 5.1. ([9]) Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu $(\kappa, \mu)$ space. Then the following identities hold:

$$
\begin{gather*}
h^{2} X=\left(\kappa+\alpha^{2}\right) \varphi^{2} X  \tag{5.1}\\
R(\xi, X) Y=\kappa(g(X, Y) \xi-\eta(Y) X)+\mu(g(X, h Y) \xi-\eta(Y) h X), \\
Q \xi=-2 n k \xi \\
\left(\nabla_{X} \varphi\right) Y=g(\alpha \varphi X+h X, Y)-\eta(Y)(\alpha \varphi X+h X)
\end{gather*}
$$

Theorem 5.1. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold. Let us suppose that $\xi \in \mathcal{N}(\kappa, \mu)$. Then, $\kappa=-1, h=0$ and $M^{2 n+1}$ is locally a warped product of an almost paraKähler manifold and an open interval. Moreover, assuming the local symmetry, $M^{2 n+1}$ is locally isometric to the hyperbolic space $H^{2 n+1}\left(-\alpha^{2}\right)$ of constant curvature $-\alpha^{2}$.

Proof. $\xi \in \mathcal{N}(\kappa, \mu)$ means that $R(X, \xi) \xi=\kappa X+\mu h X$, for any unit vector field $X$ orthogonal to $\xi$. Combining with (2.5), it follows that $h^{2} X=\left(\alpha^{2}+\kappa\right) X$. Now, if $X$ is a unit eigenvector of $h$ with eigenvalue $\lambda$, we get $\lambda^{2}=\alpha^{2}+\kappa \geq 0$. It follows that $\kappa \geq-\alpha^{2}$ and $\operatorname{Spec}(h)=\{0, \lambda,-\lambda\}$. Computing $R(X, \xi) \xi$ by means of (2.6), we easily obtain

$$
R(X, \xi) \xi=-\alpha^{2} X-2 \alpha \lambda \varphi X+\lambda^{2} X-\lambda \varphi \nabla_{\xi} X+\varphi h \nabla_{\xi} X
$$

thus we have

$$
\left(\kappa+\lambda \mu+\alpha^{2}-\lambda^{2}\right) X+2 \alpha \lambda \varphi X+\lambda \varphi \nabla_{\xi} X-\varphi h \nabla_{\xi} X=0
$$

and taking the scalar product with $\varphi X$, we obtain $\alpha \lambda=0$. Since $\alpha=$ const. $\neq 0$, it follows that $\lambda=0, h=0, \kappa=-\alpha^{2}$ and thus $K(X, \xi)=-\alpha^{2}$.

Being $h=0$, Theorem 2.2 ensures that $M^{2 n+1}$ is locally a warped product of an almost para-Kähler manifold and an open interval. Furthermore, if $M^{2 n+1}$ is locally symmetric, by Theorem 4.1, it is an $\alpha$-para-Kenmotsu manifold locally isometric to $H^{2 n+1}\left(-\alpha^{2}\right)$. Thus, we complete the proof.
From Theorem 5.1 we know for almost $\alpha$-para-Kenmotsu manifold ( $M^{2 n+1}, \varphi, \xi, \eta, g$ ), if $\xi \in \mathcal{N}(\kappa, \mu)$, then $\kappa=-1, h=0$ and $M^{2 n+1}$ is locally a warped product of an almost para-Kähler manifold and an open interval. Therefore, we consider the $(\kappa, \mu)^{\prime}$-nullity distribution, $(\kappa, \mu)^{\prime} \in R^{2}$, as the distribution $\mathcal{N}(\kappa, \mu)^{\prime}$ is given by putting for each $p \in M^{2 n+1}$,

$$
\begin{align*}
\mathcal{N}_{p}(\kappa, \mu)^{\prime} & =\left\{Z \in \Gamma\left(T_{p} M^{2 n+1}\right) \mid R(X, Y) Z\right. \\
& \left.=\kappa(g(Y, Z) X-g(X, Z) Y)+\mu\left(g(Y, Z) h^{\prime} X-g(X, Z) h^{\prime} Y\right)\right\} \tag{5.5}
\end{align*}
$$

So, if $\xi \in \mathcal{N}(\kappa, \mu)^{\prime}$, that is, for any $\mathrm{X}, \mathrm{Y} \in \Gamma\left(T M^{2 n+1}\right)$

$$
\begin{equation*}
R(X, Y) \xi=\kappa(\eta(Y) X-\eta(X) Y)+\mu\left(\eta(Y) h^{\prime} X-\eta(X) h^{\prime} Y\right) \tag{5.6}
\end{equation*}
$$

Theorem 5.2. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, \mu)^{\prime}$ and $h^{\prime} \neq 0$. Then, $\kappa<-\alpha^{2}, \mu=2 \alpha$ and $\operatorname{Spec}\left(h^{\prime}\right)=$ $\{0, \lambda,-\lambda\}$, with 0 as simple eigenvalue and $\lambda=\sqrt{-\left(\alpha^{2}+\kappa\right)}$. The distributions $[\xi] \oplus[\lambda]^{\prime}$ and $[\xi] \oplus[-\lambda]^{\prime}$ are integrable with totally geodesic leaves. The distributions $[\lambda]^{\prime}$ and $[-\lambda]^{\prime}$ are integrable with totally umbilical leaves.

Proof. Let $X$ be a unit vector field orthogonal to $\xi$, we have $R(X, \xi) \xi=k X+\mu h^{\prime} X$ and if we suppose $X \in[\lambda]^{\prime}$, since $h^{\prime 2}=-h^{2}$, combing with (2.5), we get $\lambda^{2}=$ $-\left(\kappa+\alpha^{2}\right) \geq 0$, then $\kappa \leq-\alpha^{2}$. Spec $\left(h^{\prime}\right)=\{0, \lambda,-\lambda\}$. Using (2.6) to compute $R(X, \xi) \xi$, we have

$$
\begin{equation*}
\left(\kappa+\lambda \mu+\alpha^{2}-2 \alpha \lambda+\lambda^{2}\right) X-\lambda \nabla_{\xi} X+h^{\prime} \nabla_{\xi} X=0 \tag{5.7}
\end{equation*}
$$

let (5.7) take the scalar product with $X$ and $\varphi X$ respectively, we get $\lambda(\mu-2 \alpha)=0$ and $\lambda g\left(\nabla_{\xi} X, \varphi X\right)=0$. If $\lambda=0$, then $h^{\prime}=0$ or equivalently $h=0, N(\kappa, \mu)=$
$N(\kappa, \mu)^{\prime}$ and Theorem 5.1 applies. Therefore, assuming $\lambda \neq 0$, it follows that $\kappa<-\alpha^{2}$ and $\mu=2 \alpha, g\left(\nabla_{\xi} X, \varphi X\right)=0$ for any unit $X \in[\lambda]^{\prime}$. Let (5.7) take the scalar product with any $Y \in[-\lambda]^{\prime}$, we get $g\left(\nabla_{\xi} X, Y\right)=0$ and thus $\nabla_{\xi} X \in[\lambda]^{\prime}$. Analogously $\nabla_{\xi} Y \in[-\lambda]^{\prime}$ and we obtain $\nabla_{\xi} h^{\prime}=0$. Comparing (5.6) with (2.6) for any $X, Y \in \mathcal{D}$, we have

$$
\begin{equation*}
\left(\nabla_{X} h^{\prime}\right) Y-\left(\nabla_{Y} h^{\prime}\right) X=0 \tag{5.8}
\end{equation*}
$$

If $X \in[\lambda]^{\prime}$, by (2.3) we have $\nabla_{X} \xi=\alpha X-h^{\prime} X=(\alpha-\lambda) X \in[\lambda]^{\prime}$, and since $\nabla_{\xi} h^{\prime}=0$, we easily get $\nabla_{\xi} X \in[\lambda]^{\prime}$. By (5.8) we have

$$
\begin{equation*}
0=\left(\nabla_{X} h^{\prime}\right) Z-\left(\nabla_{Z} h^{\prime}\right) X=-\lambda \nabla_{X} Z-h^{\prime} \nabla_{X} Z-\lambda \nabla_{Z} X+h^{\prime} \nabla_{Z} X \tag{5.9}
\end{equation*}
$$

let (5.9) take the scalar product with $Y \in[-\lambda]^{\prime}$, we get $g\left(\nabla_{Z} X, Y\right)=0$, therefore $\nabla_{Z} X \in[\lambda]^{\prime}$ since $g\left(\nabla_{X} Z, \xi\right)=0$. For any $X, W \in[\lambda]^{\prime}, Y, Z \in[-\lambda]^{\prime}$ it follows that $\nabla_{X} W \in[\xi] \oplus[\lambda]^{\prime}$ since $g\left(\nabla_{X} W, \xi\right)=(\lambda-\alpha) g(X, W)$. Hence, we get $g([X, W], \xi)=$ $g\left(\nabla_{X} W-\nabla_{W} X, \xi\right)=0$ and $g([X, W], Y)=g\left(\nabla_{X} W-\nabla_{W} X, Y\right)=0$, thus $[X, W] \in$ $[\lambda]^{\prime}$. Similarly, it holds $[Y, Z] \in[-\lambda]^{\prime}$. Therefore, the distributions $[\xi] \oplus[\lambda]^{\prime},[\xi] \oplus$ $[-\lambda]^{\prime},[\lambda]^{\prime}$ and $[-\lambda]^{\prime}$ are integrable. It is easy to see that the distributions $[\xi] \oplus[\lambda]^{\prime}$ and $[\xi] \oplus[-\lambda]^{\prime}$ are totally geodesic leaves. Now we prove the distribution $[\lambda]^{\prime}$ is totally umbilical, we choose a local orthonormal frame $\left\{\xi, e_{i}, \varphi e_{i}\right\}$, with $e_{i} \in[\lambda]^{\prime}$. The second fundamental form $h\left(e_{i}, e_{j}\right)=g\left(\nabla_{e_{i}} e_{j}, \xi\right) \xi=(\lambda-\alpha) \delta_{i j} \xi$, so the mean curvature vector field is $H=(\lambda-\alpha) \xi$, hence $h(X, W)=g(X, W) H$ and thus [ $\lambda]^{\prime}$ is totally umbilical. Similarly, we can get $[-\lambda]^{\prime}$ is also totally umbilical with the mean curvature vector field is $H^{\prime}=(\lambda+\alpha) \xi$ and $h^{\prime}(Y, Z)=g(Y, Z) H^{\prime}$. Thus, we complete the proof.

Theorem 5.3. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, \mu)^{\prime}$ and $h^{\prime} \neq 0$. Then, the integral manifolds of $\mathcal{D}$ are paraKähler manifolds.

Proof. For any $X, Y, Z \in \mathcal{D}$, if $\xi \in \mathcal{N}(\kappa, \mu)^{\prime}$, then $R(X, Y) \xi=0,(2.7)$ in Proposition 2.2 gives that $\left(\nabla_{h X} \Phi\right)(Y, Z)=0$. Replacing $X$ by $h X$, we get $\left(\nabla_{h^{2} X} \Phi\right)(Y, Z)=$ 0 or equivalently, $-\lambda^{2}\left(\nabla_{X} \Phi\right)(Y, Z)=0$ since $h^{2} X=-h^{\prime 2} X=-\lambda^{2} X$ if $X$ is a unit eigenvector of $h^{\prime}$ with eigenvalue $\lambda$. Being $\lambda \neq 0$, we get $\left(\nabla_{X} \Phi\right)(Y, Z)=0$. Using (2.2) we obtain $g(N(Y, Z), \varphi X)=0$, which together with (2.8) in Proposition 2.3 gives $\mathcal{N}(Y, Z)=0$ for any $Y, Z \in \mathcal{D}$, therefore the integral manifolds of $\mathcal{D}$ are para-Kähler. Thus, we complete the proof.

Corollary 5.1. Any almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, \mu)^{\prime}$, $\kappa<-\alpha^{2}$, is a $\mathcal{C R}$-manifold.

Theorem 5.4. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be a locally symmetric almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, \mu)^{\prime}$ and $h^{\prime} \neq 0$. Then, $M^{2 n+1}$ is locally isometric to $H^{n+1}\left(-(\lambda-\alpha)^{2}\right) \times R^{n}$.

Proof. As proved in Theorem 5.2, the distributions $[\xi] \oplus[\lambda]^{\prime}$ and $[\xi] \oplus[-\lambda]^{\prime}$ are integrable with totally geodesic leaves and the distributions $[\lambda]^{\prime}$ and $[-\lambda]^{\prime}$ are integrable with totally umbilical leaves. It follows that $M^{2 n+1}$ is locally isometric to the product of an integral manifold $M_{1}^{n+1}$ of $[\xi] \oplus[\lambda]^{\prime}$ and an integral manifold $M_{2}^{n}$ of $[-\lambda]^{\prime}$. Therefore, we can choose coordinates $\left(u^{0}, \ldots, u^{2 n}\right)$ such that $\frac{\partial}{\partial u^{0}} \in[\xi], \frac{\partial}{\partial u^{1}}, \ldots, \frac{\partial}{\partial u^{n}} \in[\lambda]^{\prime}$ and $\frac{\partial}{\partial u^{n+1}}, \ldots, \frac{\partial}{\partial u^{2 n}} \in[-\lambda]^{\prime}$. Now, we set $X_{i}=\frac{\partial}{\partial u^{2}}$ for any $i \in\{1, \ldots, n\}$, so that the distribution $[-\lambda]^{\prime}$ is spanned by the vector fields $\varphi X_{1}, \ldots, \varphi X_{n}$. it is easy to see that $X_{i} \in[\lambda]^{\prime}$ is projectable and $\varphi X_{i} \in[-\lambda]^{\prime}$ is vertical, then $\left[X_{i}, \varphi X_{j}\right]$ is vertical by $[1]$, hence $\left[X_{i}, \varphi X_{j}\right] \in[-\lambda]^{\prime}$. Taking the scalar product with any $Z \in[\lambda]^{\prime}$, since $\nabla_{X_{i}} \varphi X_{j} \in[-\lambda]^{\prime}$, we get $g\left(\nabla_{\varphi X_{j}} X_{i}, Z\right)=0$ and then $\nabla_{\varphi X_{j}} X_{i}=0$. Applying $\left(\nabla_{\varphi X} \varphi\right) \varphi Y-\left(\nabla_{X} \varphi\right) Y=\alpha(\eta(Y) \varphi X+2 g(X, \varphi Y) \xi)+$ $\eta(Y) h X$ (appeared in [9]), we have $\left(\nabla_{X_{i}} \varphi\right) X_{j}+\varphi\left(\nabla_{\varphi X_{i}} \varphi X_{j}\right)=0$, which implies $\left(\nabla_{X_{i}} \varphi\right) X_{j}=0, \nabla_{\varphi X_{i}} \varphi X_{j}=0$, since the two part belong to $[-\lambda]^{\prime}$ and $[\lambda]^{\prime}$ respectively. $\nabla_{\varphi X_{i}} \varphi X_{j}=0$ means that $M_{2}^{n}$ of $[-\lambda]^{\prime}$ is flat. Now we compute the curvature of $M_{1}^{n+1}$. Applying $\varphi$ to $\left(\nabla_{X_{i}} \varphi\right) X_{j}=0$ gives

$$
\nabla_{X_{i}} X_{j}-\varphi \nabla_{X_{i}} \varphi X_{j}=(\lambda-\alpha) g\left(X_{i}, X_{j}\right) \xi
$$

Derivating with respect to $X_{k}$ yields:

$$
\begin{aligned}
& \nabla_{X_{k}} \nabla_{X_{i}} X_{j}-\left(\nabla_{X_{k}} \varphi\right)\left(\nabla_{X_{i}} \varphi X_{j}\right)-\varphi \nabla_{X_{k}} \nabla_{X_{i}} \varphi X_{j} \\
= & (\lambda-\alpha) X_{k}\left(g\left(X_{i}, X_{j}\right)\right) \xi-(\lambda-\alpha)^{2} g\left(X_{i}, X_{j}\right) X_{k} .
\end{aligned}
$$

taking the scalar product with $X_{l}$ on both sides of the above equality and taking into account $g\left(\left(\nabla_{X_{k}} \varphi\right)\left(\nabla_{X_{i}} \varphi X_{j}\right), X_{l}\right)=-g\left(\nabla_{X_{i}} \varphi X_{j},\left(\nabla_{X_{k}} \varphi\right) X_{l}\right)=0$, we obtain

$$
g\left(\nabla_{X_{k}} \nabla_{X_{i}} X_{j}, X_{l}\right)+g\left(\nabla_{X_{k}} \nabla_{X_{i}} \varphi X_{j}, \varphi X_{l}\right)=-(\lambda-\alpha)^{2} g\left(X_{i}, X_{j}\right) g\left(X_{k}, X_{l}\right) .
$$

Interchanging $i$ and $k$, subtracting and being $\left[X_{i}, X_{k}\right]=0$ we have

$$
\begin{aligned}
& g\left(R\left(X_{k}, X_{i}\right) X_{j}, X_{l}\right)+g\left(R\left(X_{k}, X_{i}\right) \varphi X_{j}, \varphi X_{l}\right) \\
=\quad & -(\lambda-\alpha)^{2} g\left(X_{i}, X_{j}\right) g\left(X_{k}, X_{l}\right)+(\lambda-\alpha)^{2} g\left(X_{k}, X_{j}\right) g\left(X_{i}, X_{l}\right)
\end{aligned}
$$

Since $\nabla_{\varphi X_{i}} \varphi X_{j}=0$ and $\left[\varphi X_{i}, \varphi X_{j}\right]=0$, by a straightforward calculation we obtain

$$
g\left(R\left(X_{k}, X_{i}\right) \varphi X_{j}, \varphi X_{l}\right)=g\left(R\left(\varphi X_{j}, \varphi X_{l}\right) X_{k}, X_{i}\right)=0
$$

and thus

$$
g\left(R\left(X_{k}, X_{i}\right) X_{j}, X_{l}\right)=-(\lambda-\alpha)^{2}\left[g\left(X_{i}, X_{j}\right) g\left(X_{k}, X_{l}\right)-g\left(X_{k}, X_{j}\right) g\left(X_{i}, X_{l}\right)\right]
$$

Moreover, since $R(X, Y) \xi=0$ for any $X, Y \in \mathcal{D}$, we get $g\left(R\left(X_{i}, X_{j}\right) \xi, X_{k}\right)=$ 0 . By (2.4) in Proposition 2.2, and $\nabla_{\xi} h=0$ because of the symmetry, we get $g\left(R\left(X_{i}, \xi\right) \xi, X_{j}\right)=-(\lambda-\alpha)^{2} g\left(X_{i}, X_{j}\right)$. Therefore, we conclude that the integral manifold $M_{1}^{n+1}$ of $[\xi] \oplus[\lambda]^{\prime}$ is a space of constant curvature $-(\lambda-\alpha)^{2}$. Thus, we complete the proof.

Lemma 5.1. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in N(\kappa, 2 \alpha)^{\prime}$ and $h^{\prime} \neq 0$. Then, for any $X, Y \in \Gamma\left(T M^{2 n+1}\right)$,

$$
\begin{equation*}
\left(\nabla_{X} h^{\prime}\right) Y=g\left(h^{\prime 2} X-\alpha h^{\prime} X, Y\right) \xi+\eta(Y)\left(h^{\prime 2} X-\alpha h^{\prime} X\right) \tag{5.10}
\end{equation*}
$$

Proof. We choose a local orthonormal frame $\left\{\xi, e_{i}, \varphi e_{i}\right\}$ with $e_{i} \in[\lambda]^{\prime}$.

1) If $X, Y \in[\lambda]^{\prime}$, we know that $\nabla_{X} Y \in[\xi] \oplus[\lambda]^{\prime}$ from Theorem 5.2. It is easy to get

$$
\nabla_{X} Y=g\left(\nabla_{X} Y, e_{i}\right) e_{i}+g\left(\nabla_{X} Y, \xi\right) \xi=(\lambda-\alpha) g(X, Y) \xi+g\left(\nabla_{X} Y, e_{i}\right) e_{i}
$$

and thus

$$
\left(\nabla_{X} h^{\prime}\right) Y=\nabla_{X} h^{\prime} Y-h^{\prime} \nabla_{X} Y=\lambda \nabla_{X} Y-\lambda g\left(\nabla_{X} Y, e_{i}\right) e_{i}=\lambda(\lambda-\alpha) g(X, Y) \xi
$$

2) If $X, Y \in[-\lambda]^{\prime}$, we know that $\nabla_{X} Y \in[\xi] \oplus[-\lambda]^{\prime}$ from Theorem 5.2. Similarly we have

$$
\nabla_{X} Y=g\left(\nabla_{X} Y, \varphi e_{i}\right) \varphi e_{i}+g\left(\nabla_{X} Y, \xi\right) \xi=-(\lambda+\alpha) g(X, Y) \xi+g\left(\nabla_{X} Y, \varphi e_{i}\right) \varphi e_{i}
$$

and

$$
\left(\nabla_{X} h^{\prime}\right) Y=\lambda(\lambda+\alpha) g(X, Y) \xi
$$

3) If $X \in[\lambda]^{\prime}, Y \in[-\lambda]^{\prime}$, since $g\left(\nabla_{X} Y, \xi\right)=(\lambda-\alpha) g(X, Y)=0$, and for any $Z \in[\lambda]^{\prime}, g\left(\nabla_{X} Y, Z\right)=X g(Y, Z)-g\left(Y, \nabla_{X} Z\right)=0$, thus we get $\nabla_{X} Y \in[-\lambda]^{\prime}$ and $\left(\nabla_{X} h^{\prime}\right) Y=\nabla_{X} h^{\prime} Y-h^{\prime} \nabla_{X} Y=0$, therefore we have $\left(\nabla_{Y} h^{\prime}\right) X=0$ since $\left(\nabla_{X} h^{\prime}\right) Y-\left(\nabla_{Y} h^{\prime}\right) X=0$.

Therefore, for any $X \in \Gamma\left(T M^{2 n+1}\right)$, we write $X=X_{\lambda}+X_{-\lambda}+\eta(X) \xi$, with $X_{\lambda} \in[\lambda]^{\prime}$ and $X_{-\lambda} \in[-\lambda]^{\prime}$, since $\nabla_{\xi} h^{\prime}=0$, we get

$$
\begin{aligned}
\left(\nabla_{X} h^{\prime}\right) Y= & \left(\nabla_{X_{\lambda}} h^{\prime}\right) Y_{\lambda}+\eta(Y)\left(\nabla_{X_{\lambda}} h^{\prime}\right) \xi+\left(\nabla_{X_{-\lambda}} h^{\prime}\right) Y_{-\lambda}+\eta(Y)\left(\nabla_{X_{-\lambda}} h^{\prime}\right) \xi \\
= & \lambda(\lambda-\alpha) g\left(X_{\lambda}, Y_{\lambda}\right) \xi+\lambda(\lambda-\alpha) \eta(Y) X_{\lambda}+\lambda(\lambda+\alpha) g\left(X_{-\lambda}, Y_{-\lambda}\right) \xi \\
& +\lambda(\lambda+\alpha) \eta(Y) X_{-\lambda} \\
= & -\alpha \lambda\left\{g\left(X_{\lambda}, Y_{\lambda}\right)-g\left(X_{-\lambda}, Y_{-\lambda}\right)\right\} \xi+\lambda^{2}\left\{g\left(X_{\lambda}, Y_{\lambda}\right)+g\left(X_{-\lambda}, Y_{-\lambda}\right)\right\} \xi \\
& +\eta(Y)\left(-\alpha \lambda X_{\lambda}+\alpha \lambda X_{-\lambda}+\lambda^{2} X_{\lambda}-\lambda^{2} X_{-\lambda}\right) \\
= & g\left(h^{\prime 2} X-\alpha h^{\prime} X, Y\right) \xi+\eta(Y)\left(h^{\prime 2} X-\alpha h^{\prime} X\right) .
\end{aligned}
$$

Lemma 5.2. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$ and $h^{\prime} \neq 0$. Then, for any $X, Y \in \mathcal{D}$,
$R(X, Y) h^{\prime} Z-h^{\prime} R(X, Y) Z=\left(\kappa+2 \alpha^{2}\right)\left[g(Y, Z) h^{\prime} X-g(X, Z) h^{\prime} Y-g\left(h^{\prime} Y, Z\right) X+g\left(h^{\prime} X, Z\right) Y\right]$.

Proof. We know from Lemma 5.1 that $\left(\nabla_{X} h^{\prime}\right) Y=g\left(h^{\prime 2} X-\alpha h^{\prime} X, Y\right) \xi$ for any $X, Y \in \mathcal{D}$, by direct calculation we obtain

$$
\begin{aligned}
& R(X, Y) h^{\prime} Z-h^{\prime} R(X, Y) Z \\
& =\nabla_{X} \nabla_{Y} h^{\prime} Z-\nabla_{Y} \nabla_{X} h^{\prime} Z-\nabla_{[X, Y]} h^{\prime} Z-h^{\prime} R(X, Y) Z \\
& =g\left(\left(\nabla_{X} h^{\prime 2}\right) Y-\left(\nabla_{Y} h^{\prime 2}\right) X-\alpha\left(\left(\nabla_{X} h^{\prime}\right) Y-\left(\nabla_{Y} h^{\prime}\right) X\right), Z\right)+g\left(h^{\prime 2} Y-\alpha h^{\prime} Y, Z\right) \nabla_{X} \xi \\
& -g\left(h^{\prime 2} X-\alpha h^{\prime} X, Z\right) \nabla_{Y} \xi-g\left(\nabla_{Y} \xi, Z\right)\left(h^{\prime 2} X-\alpha h^{\prime} X\right)+g\left(\nabla_{X} \xi, Z\right)\left(h^{\prime 2} Y-\alpha h^{\prime} Y\right) .
\end{aligned}
$$

It follows that for any $X, Y \in \mathcal{D}$, we know from $h^{\prime 2} X=-h^{2} X=-\left(\kappa+\alpha^{2}\right) X$, and $\left(\nabla_{X} h^{\prime 2}\right) Y=-\left(\kappa+\alpha^{2}\right) \eta\left(\nabla_{X} Y\right) \xi$, hence, $\left(\nabla_{X} h^{2}\right) Y-\left(\nabla_{Y} h^{\prime 2}\right) X=0$ since $\mathcal{D}$ is integrable, and from Lemma 5.1, we get $\left(\nabla_{X} h^{\prime}\right) Y-\left(\nabla_{Y} h^{\prime}\right) X=0$. Lemma 5.2 is followed by direct computation. Thus, we complete the proof.

Lemma 5.3. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$ and $h^{\prime} \neq 0$. Then, for any $X, Y, Z \in \mathcal{D}$, we have

$$
\begin{aligned}
& R(X, Y) \varphi Z-\varphi R(X, Y) Z \\
= & g\left(\alpha X-h^{\prime} X, \varphi Z\right)\left(\alpha Y-h^{\prime} Y\right)-g\left(\alpha X-h^{\prime} X, Z\right)\left(\alpha \varphi Y-\varphi h^{\prime} Y\right) \\
& +g\left(\alpha Y-h^{\prime} Y, Z\right)\left(\alpha \varphi X-\varphi h^{\prime} X\right)-g\left(\alpha Y-h^{\prime} Y, \varphi Z\right)\left(\alpha X-h^{\prime} X\right) .
\end{aligned}
$$

Proof. Since the Weingarten operator for an integral manifold $M^{\prime}$ of $\mathcal{D}$ is given by

$$
A X=-\nabla_{X} \xi=-\left(\alpha X-h^{\prime} X\right)
$$

by Theorem 2.3 in [4] we get the Guass equation
$R(X, Y) Z=R^{\prime}(X, Y) Z+g\left(\alpha X-h^{\prime} X, Z\right)\left(\alpha Y-h^{\prime} Y\right)-g\left(\alpha Y-h^{\prime} Y, Z\right)\left(\alpha X-h^{\prime} X\right)$.
By Theorem 5.3, the integral manifolds of $\mathcal{D}$ are para-Kähler manifolds, and from Lemma 10.1 of [4], we know $R^{\prime}(X, Y) \varphi Z-\varphi R^{\prime}(X, Y) Z=0$. Combining with the above two equations, we get the required formula for $R$ and $\varphi$. Thus, we complete the proof.

Proposition 5.2. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$ and $h^{\prime} \neq 0$. Then, for any $X_{\lambda}, Y_{\lambda}, Z_{\lambda} \in[\lambda]^{\prime}$ and $X_{-\lambda}, Y_{-\lambda}, Z_{-\lambda} \in[-\lambda]^{\prime}$, the curvature tensor $R$ satisfies:

$$
\begin{gathered}
R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}=0, \\
R\left(X_{-\lambda}, Y_{-\lambda}\right) Z_{\lambda}=0, \\
R\left(X_{\lambda}, Y_{-\lambda}\right) Z_{\lambda}=\left(\kappa+2 \alpha^{2}\right) g\left(X_{\lambda}, Z_{\lambda}\right) Y_{-\lambda}, \\
R\left(X_{\lambda}, Y_{-\lambda}\right) Z_{-\lambda}=-\left(\kappa+2 \alpha^{2}\right) g\left(Y_{-\lambda}, Z_{-\lambda}\right) X_{\lambda}, \\
R\left(X_{\lambda}, Y_{\lambda}\right) Z_{\lambda}=(\kappa+2 \alpha \lambda)\left[g\left(Y_{\lambda}, Z_{\lambda}\right) X_{\lambda}-g\left(X_{\lambda}, Z_{\lambda}\right) Y_{\lambda}\right], \\
R\left(X_{-\lambda}, Y_{-\lambda}\right) Z_{-\lambda}=(\kappa-2 \alpha \lambda)\left[g\left(Y_{-\lambda}, Z_{-\lambda}\right) X_{-\lambda}-g\left(X_{-\lambda}, Z_{-\lambda}\right) Y_{-\lambda}\right] .
\end{gathered}
$$

Proof. For any $X \in[\lambda]^{\prime}$, and $Y, Z \in[-\lambda]^{\prime}$, by Lemma 5.2 we have

$$
-\lambda R(X, Y) Z-h^{\prime} R(X, Y) Z=2 \lambda\left(\kappa+2 \alpha^{2}\right) g(Y, Z) X
$$

Taking the scalar product with $W \in[\lambda]^{\prime}$, we obtain

$$
\begin{equation*}
g(R(X, Y) Z, W)=-\left(\kappa+2 \alpha^{2}\right) g(Y, Z) g(X, W) \tag{5.11}
\end{equation*}
$$

Lemma 5.2 implies that $R(X, Y) Z \in[\lambda]^{\prime}$ for any $X, Y, Z \in[\lambda]^{\prime}$ and $R(X, Y) Z \in$ $[-\lambda]^{\prime}$ for any $X, Y, Z \in[-\lambda]^{\prime}$. Now, in order to compute $R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}$, we consider a local orthonormal frame $\left\{\xi, e_{i}, \varphi e_{i}\right\}$, with $e_{i} \in[\lambda]^{\prime}$. Condition $\xi \in N(\kappa, 2 \alpha)^{\prime}$ means that $g\left(R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}, \xi\right)=g\left(R\left(X_{\lambda}, Y_{\lambda}\right) \xi, Z_{-\lambda}\right)=0$, and since $R\left(X_{\lambda}, Y_{\lambda}\right) e_{i} \in[\lambda]^{\prime}$, thus $g\left(R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}, e_{i}\right)=0$. Using the first Bianchi identity and (5.11), we have

$$
\begin{aligned}
g\left(R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}, \varphi e_{i}\right) & =g\left(R\left(Y_{\lambda}, Z_{-\lambda}\right) \varphi e_{i}, X_{\lambda}\right)-g\left(R\left(X_{\lambda}, Z-\lambda\right) \varphi e_{i}, Y_{\lambda}\right) \\
& =-\left(\kappa+2 \alpha^{2}\right)\left[g\left(Z_{-\lambda}, \varphi e_{i}\right) g\left(X_{\lambda}, Y_{\lambda}\right)-g\left(Z_{-\lambda}, \varphi e_{i}\right) g\left(X_{\lambda}, Y_{\lambda}\right)\right] \\
& =0
\end{aligned}
$$

so that $R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}=0$. The terms $R\left(X_{-\lambda}, Y_{-\lambda}\right) Z_{\lambda}, R\left(X_{\lambda}, Y_{-\lambda}\right) Z_{\lambda}$ and $R\left(X_{\lambda}, Y_{-\lambda}\right) Z_{-\lambda}$ are computed in a similar manner. By Lemma 5.3, using $R\left(X_{\lambda}, Y_{\lambda}\right) Z_{-\lambda}=0$, we get

$$
R\left(X_{\lambda}, Y_{\lambda}\right) \varphi Z_{\lambda}=-(\alpha-\lambda)^{2}\left[g\left(Y_{\lambda}, \varphi Z_{-\lambda}\right) X_{\lambda}-g\left(X_{\lambda}, \varphi Z_{-\lambda}\right) Y_{\lambda}\right]
$$

Replacing $Z_{-\lambda}$ by $\varphi Z_{\lambda} \in[\lambda]^{\prime}$, and since $-(\alpha-\lambda)^{2}=\kappa+2 \alpha \lambda$, we have

$$
R\left(X_{\lambda}, Y_{\lambda}\right) Z_{\lambda}=R\left(X_{\lambda}, Y_{\lambda}\right) \varphi\left(\varphi Z_{\lambda}\right)=(\kappa+2 \alpha \lambda)\left[g\left(Y_{\lambda}, Z_{\lambda}\right) X_{\lambda}-g\left(X_{\lambda}, Z_{\lambda}\right) Y_{\lambda}\right] .
$$

In the same manner, we obtain $R\left(X_{-\lambda}, Y_{-\lambda}\right) Z_{-\lambda}=(\kappa-2 \alpha \lambda)\left[g\left(Y_{-\lambda}, Z_{-\lambda}\right) X_{-\lambda}-\right.$ $\left.g\left(X_{-\lambda}, Z_{-\lambda}\right) Y_{-\lambda}\right]$. Thus, we complete the proof.

Proposition 5.3. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$ and $h^{\prime} \neq 0$. Then, we have

1) $K(X, \xi)=\kappa+2 \alpha \lambda$, if $X \in[\lambda]^{\prime}$; $K(X, \xi)=\kappa-2 \alpha \lambda$, if $X \in[-\lambda]^{\prime} ;$
2) $K(X, Y)=\kappa+2 \alpha \lambda$, if $X, Y \in[\lambda]^{\prime}$; $K(X, Y)=\kappa-2 \alpha \lambda$, if $X, Y \in[-\lambda]^{\prime} ;$ $K(X, Y)=-\left(\kappa+2 \alpha^{2}\right)$, if $X \in[\lambda]^{\prime}, Y \in[-\lambda]^{\prime}$.
3) $r=8 \alpha \lambda n-4 \alpha^{2} n^{2}-2 k n$.

Proof. The proof for the sectional curvature is easily followed by Proposition 5.2. In order to compute the scalar curvature, we choose a orthonormal frame $\left\{\xi, e_{i}, \varphi e_{i}\right\}$ with $e_{i} \in[\lambda]^{\prime}$, by direct calculations we have

$$
\operatorname{Ric}(\xi, \xi)=\sum_{i=1}^{n} R\left(\xi, e_{i}, e_{i}, \xi\right)-\sum_{i=1}^{n} R\left(\xi, \varphi e_{i}, \varphi e_{i}, \xi\right)=4 \alpha \lambda n
$$

$$
\begin{aligned}
\operatorname{Ric}\left(e_{i}, e_{i}\right)= & \sum_{i=1}^{n} R\left(e_{i}, \xi, \xi, e_{i}\right)+\sum_{j \neq i=1}^{n} R\left(e_{i}, e_{j}, e_{j}, e_{i}\right)-\sum_{j=1}^{n} R\left(e_{i}, \varphi e_{j}, \varphi e_{j}, e_{i}\right) \\
= & n(\kappa+2 \alpha \lambda)+n\left(\kappa+2 \alpha^{2}\right), \\
& \operatorname{Ric}\left(\varphi e_{i}, \varphi e_{i}\right)=(\kappa-2 \alpha \lambda)(2-n)+n\left(\kappa+2 \alpha^{2}\right),
\end{aligned}
$$

and it is easy to get the scalar curvature $r=8 \alpha \lambda n-4 \alpha^{2} n^{2}-2 k n$.
Proposition 5.4. Let $\left(M^{2 n+1}, \varphi, \xi, \eta, g\right)$ be an almost $\alpha$-para-Kenmotsu manifold such that $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$ and $h^{\prime} \neq 0$. Then, $M^{2 n+1}$ is locally isometric to the warped products

$$
S^{n+1}(\kappa+2 \alpha \lambda) \times_{f} R^{n}, \quad \text { or } \quad B^{n+1}(\kappa-2 \alpha \lambda) \times_{f^{\prime}} R^{n}
$$

where $S^{n+1}(\kappa+2 \alpha \lambda)$ is a space of constant positive curvature $\kappa+2 \alpha \lambda, B^{n+1}(\kappa-2 \alpha \lambda)$ is a space of constant negative curvature $\kappa-2 \alpha \lambda, f=c e^{-(\lambda+\alpha) t}, f^{\prime}=c^{\prime} e^{(\alpha-\lambda) t}$, with $c, c^{\prime}$ positive constants.

Proof. By Theorem 5.2, we get that the distributions $[\xi] \oplus[\lambda]^{\prime}$ and $[\xi] \oplus[-\lambda]^{\prime}$ are integrable with totally geodesic leaves, the distributions $[\lambda]^{\prime}$ and $[-\lambda]^{\prime}$ are integrable with totally umbilical leaves. First, we consider that $M^{2 n+1}$ is locally a warped product $S \times_{f} F$ such that $T S=[\xi] \oplus[\lambda]^{\prime}$ and $T F=[-\lambda]^{\prime}$. Now, we compute the function $f$. We have denoted by $\check{g}$ and $\hat{g}$ the pseudo-Riemannian metrics on $S$ and $F$, respectively, such that the warped metric is given by $\check{g}+f^{2} \hat{g}$. Then, the projection $\pi: S \times_{f} F \rightarrow S$ is a submersion with horizontal distribution $[\xi] \oplus[\lambda]^{\prime}$ and vertical distribution $[-\lambda]^{\prime}$. From Theorem 5.2 we know that the mean curvature vector field for the immersed submanifold $(F, \hat{g})$ is $H^{\prime}=(\lambda+\alpha) \xi$. By Proposition 4.1 in [4], we get for any $Y, Z \in[-\lambda]^{\prime}, \operatorname{nor}\left(\nabla_{Y} Z\right)=h(Y, Z)=-\frac{g(Y, Z)}{f} \operatorname{grad}_{\check{g}} f$. And since $h(Y, Z)=g(Y, Z) H^{\prime}$, we get $-(\lambda+\alpha) f \xi=\operatorname{grad}_{\check{g}} f$. We choose local coordinates $\left\{t, x^{1}, \ldots, x^{n}\right\}$ on $B$ such that $\xi=\frac{\partial}{\partial t}$ and $\frac{\partial}{\partial x^{i}} \in[\lambda]^{\prime}$ for any $i=$ $1, \ldots, n$. After direct computation we get $f=c e^{-(\lambda+\alpha) t}, c>0$. Since $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$, we have $R(X, Y) \xi=0$, and $R(X, \xi) \xi=(\kappa+2 \alpha \lambda) X$, also by $\xi \in \mathcal{N}(\kappa, 2 \alpha)^{\prime}$, we get $R(\xi, X) Y=\kappa(g(X, Y) \xi-\eta(Y) X)+2 \alpha\left(g\left(h^{\prime} X, Y\right) \xi-\eta(Y)^{\prime} h X\right)$, thus, we get $R(\xi, X) Y=(\kappa+2 \alpha \lambda) g(X, Y) \xi$. Applying Proposition 5.2, we get $R(X, Y) Z=$ $(\kappa+2 \alpha \lambda)[g(Y, Z) X-g(X, Z) Y]$, hence, we conclude that $S$ is a space of constant curvature $\kappa+2 \alpha \lambda>0$. Next, we compute the curvature $R^{F}$ of $(F, \hat{g})$, by Proposition 4.2 in [4], for any $U, V, W \in[-\lambda]^{\prime}$, it holds

$$
R^{F}(V, W) U=R(V, W) U-\frac{g(\operatorname{grad} f, \operatorname{grad} f)}{f^{2}}\{g(V, U) W-g(W, U) V\}
$$

Since $\operatorname{grad} f=-(\lambda+\alpha) f \xi$, we get that $g(\operatorname{grad} f, \operatorname{grad} f)=(\lambda+\alpha)^{2} f^{2}=(2 \alpha \lambda-\kappa) f^{2}$, and by Proposition 5.2, we get $R(V, W) U=(2 \alpha \lambda-\kappa)\{g(V, U) W-g(W, U) V\}$. Then, $R^{F}(V, W) U=0$, and thus the fibers of the warped product are flat spaces.

Similar discussions for horizontal distribution $[\xi] \oplus[-\lambda]^{\prime}$ and vertical distribution $[\lambda]^{\prime}$. In this case, the mean curvature vector field for the immersed submanifold
$(F, \hat{g})$ is $H^{\prime}=(\lambda-\alpha) \xi$ and computing the warping function, we obtain $f^{\prime}=$ $c^{\prime} e^{(\alpha-\lambda) t}, c^{\prime}>0$. Moreover, we can also prove that the base manifold of the warped product is a space of constant curvature $\kappa-2 \alpha \lambda<0$ and the fibers are flat spaces. Thus, we complete the proof.

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Quanxiang Pan
Henan Institute of Techonology
School of Science
453000 Xinxiang, China
panquanxiang@dlut.edu.cn

Ximin Liu<br>Dalian University of Technology<br>School of Mathematical Sciences<br>116024,Dalian, China<br>ximinliu@dlut.edu.cn

# APPROXIMATION BY JAIN-SCHURER OPERATORS 

Nursel Çetin and Gülen Başcanbaz-Tunca

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#### Abstract

In this paper we deal with Jain-Schurer operators. We give an estimate, related to the degree of approximation, via moduli of smoothness of the first and the second order. Also, we present a Voronovskaja-type result. Moreover, we show that the Jain-Schurer operators preserve the properties of a modulus of continuity. Finally, we study monotonicity of the sequence of the Jain-Schurer operators when the attached function is convex and non-decreasing.


Keywords: Jain-Schurer operators; monotonicity; moduli of smoothness; Voronovskajatype result.

## 1. Introduction

In [19], Schurer constructed the following linear positive operators

$$
\begin{equation*}
S_{n, p}(f ; x)=e^{-(n+p) x} \sum_{k=0}^{\infty} f\left(\frac{k}{n}\right) \frac{(n+p)^{k} x^{k}}{k!} \tag{1.1}
\end{equation*}
$$

where $x \in[0, b], b<\infty, n \in \mathbb{N}, p \geq 0$, and $f$ is real valued and bounded function on $[0, \infty)$. The case $p=0$ gives the the well known Szász-Mirakjan operators. There are a number of generalizations of Szász-Mirakjan operators, here we cite only a few ([4], [6], [10], [11]) with references therein. Some works concerning Schurer's setting can be found in [3], [14], [20], [16] and [17]. Motivated by these statements, we extend the well known Jain operators in the Schurer's design. Recall that in [12], Jain constructed the following linear positive operators

$$
\begin{equation*}
P_{n}^{[\beta]}(f ; x)=\sum_{k=0}^{\infty} f\left(\frac{k}{n}\right) w_{\beta}(k ; n x), \quad x \in(0, \infty), \tag{1.2}
\end{equation*}
$$

and $P_{n}^{[\beta]}(f ; 0)=f(0)$, where $n \in \mathbb{N}, \beta \in[0,1), f \in C[0, \infty)$, and for $0<\alpha<$ $\infty, w_{\beta}(k ; \alpha)$ is given by

$$
\begin{equation*}
w_{\beta}(k ; \alpha):=\frac{\alpha(\alpha+k \beta)^{k-1}}{k!} e^{-(\alpha+k \beta)}, \quad k \in \mathbb{N} \cup\{0\} \tag{1.3}
\end{equation*}
$$

and it satisfies $\sum_{k=0}^{\infty} w_{\beta}(k ; \alpha)=1$. In the paper, the author studied convergence properties and the order of approximation by the sequence of these operators on any finite closed interval of $[0, \infty)$ by taking $\beta$ as a sequence $\beta_{n}$ such that $\beta_{n} \rightarrow 0$ as $n \rightarrow \infty$. For some interesting papers concerning Jain operators, we refer to [1], [2], [7], [9], [18], [23] and references therein. Obviously, the case $\beta=0$ gives the well known Szász-Mirakjan operators [22].

In this work, for a fixed $p \in \mathbb{N} \cup\{0\}$, we consider the linear positive operators denoted by $S_{n, p}^{\beta}, n \in \mathbb{N}$, and defined as

$$
\begin{equation*}
S_{n, p}^{\beta}(f ; x)=\sum_{k=0}^{\infty} f\left(\frac{k}{n}\right) w_{\beta}(k ;(n+p) x), \quad x \in(0, \infty) \tag{1.4}
\end{equation*}
$$

and $S_{n, p}^{\beta}(f ; 0)=f(0)$, for $f \in C_{B}[0, \infty):=\{f \in C[0, \infty): f$ is bounded $\}$, $\beta \in[0,1)$, and $w_{\beta}(k ;(n+p) x)$ given by (1.3). We call $S_{n, p}^{\beta}$ as Jain-Schurer operators. Note that, each $S_{n, p}^{\beta}$ maps $C_{B}[0, \infty)$ into itself, and the case $p=0$ covers the Jain operators: $S_{n, 0}^{\beta}=P_{n}^{[\beta]}, n \in \mathbb{N}$. On the other hand, in the case $\beta=0, S_{n, p}^{\beta}$ reduces to the Schurer extension of the Szász-Mirakjan operators given by (1.1). We obtain an estimate, which will be used next for the rate of convergence, with the help of the modulus of smoothness of a bounded and continuous function, and prove a Voronovskaja-type result. Moreover, we show that each JainSchurer operator preserves the properties of a general modulus of continuity. Finally, we investigate the monotonicity of the sequence of the Jain-Schurer operators $S_{n, p}^{\beta}(f)$, with respect to $n$, when the function $f$ is convex and non-decreasing.

Now, denoting $e_{j}(t)=t^{j}, j \in \mathbb{N} \cup\{0\}$ and $\varphi_{x}^{j}(t):=(t-x)^{j}, j \in \mathbb{N}$, for the Jain operators $P_{n}^{[\beta]}$ we have (see, e.g., [11, Lemma 1])

Lemma 1.1. For the operators $P_{n}^{[\beta]}$ given by (1.2), one has

$$
\begin{aligned}
P_{n}^{[\beta]}\left(e_{0} ; x\right) & =1, \\
P_{n}^{[\beta]}\left(e_{1} ; x\right) & =\frac{x}{1-\beta}, \\
P_{n}^{[\beta]}\left(e_{2} ; x\right) & =\frac{x^{2}}{(1-\beta)^{2}}+\frac{x}{n(1-\beta)^{3}}, \\
P_{n}^{[\beta]}\left(e_{3} ; x\right) & =\frac{x^{3}}{(1-\beta)^{3}}+\frac{3 x^{2}}{n(1-\beta)^{4}}+\frac{(1+2 \beta) x}{n^{2}(1-\beta)^{5}}, \\
P_{n}^{[\beta]}\left(e_{4} ; x\right) & =\frac{x^{4}}{(1-\beta)^{4}}+\frac{6 x^{3}}{n(1-\beta)^{5}}+\frac{(8 \beta+7) x^{2}}{n^{2}(1-\beta)^{6}}+\frac{\left(6 \beta^{2}+8 \beta+1\right) x}{n^{3}(1-\beta)^{7}} .
\end{aligned}
$$

Making use of Lemma 1.1, straightforward computation shows that moments and central moments of the Jain-Schurer operators are obtained as in the following lemmas, respectively:

Lemma 1.2. For the operators $S_{n, p}^{\beta}$ given by (1.4), one has

$$
S_{n, p}^{\beta}\left(e_{j} ; x\right)=P_{n}^{[\beta]}\left(e_{j} ;\left(\frac{n+p}{n}\right) x\right), \quad j=0,1, \ldots
$$

Lemma 1.3. For the operators $S_{n, p}^{\beta}$ given by (1.4), one has

$$
\begin{aligned}
S_{n, p}^{\beta}\left(\varphi_{x}^{1} ; x\right)= & \left(\beta+\frac{p}{n}\right) \frac{x}{1-\beta}, \\
S_{n, p}^{\beta}\left(\varphi_{x}^{2} ; x\right)= & \left(\beta+\frac{p}{n}\right)^{2} \frac{x^{2}}{(1-\beta)^{2}}+\left(1+\frac{p}{n}\right) \frac{x}{n(1-\beta)^{3}}, \\
S_{n, p}^{\beta}\left(\varphi_{x}^{4} ; x\right)= & \left(\beta+\frac{p}{n}\right)^{4} \frac{x^{4}}{(1-\beta)^{4}}+6\left(\beta+\frac{p}{n}\right)^{2}\left(1+\frac{p}{n}\right) \frac{x^{3}}{n(1-\beta)^{5}} \\
& +\left(1+\frac{p}{n}\right) \frac{\left(4 n \beta+3 n+8 p \beta+7 p+8 n \beta^{2}\right)}{n^{3}(1-\beta)^{6}} x^{2}+\left(1+\frac{p}{n}\right) \frac{\left(6 \beta^{2}+8 \beta+1\right)}{n^{3}(1-\beta)^{7}} x .
\end{aligned}
$$

## 2. Modulus of smoothness $K$-Functional

In this part of the paper, we extend the result proved by Agratini for the Jain operators [2, Theorem 2] to the Jain-Schurer operators. To this aim, we recall the terminology that will be used in the results. As usual, let $C_{B}[0, \infty)$ denote the space of real valued, bounded and continuous functions defined on $[0, \infty)$ equipped with the norm given by

$$
\|f\|=\sup _{x \in[0, \infty)}|f(x)|
$$

for $f \in C_{B}[0, \infty)$. Also, let $U C_{B}[0, \infty)$ denote the space of all real valued bounded and uniformly continuous functions on $[0, \infty)$.

For a bounded, real valued function $f$ on $[0, \infty)$ and $\delta>0$, the first modulus of smoothness, modulus of continuity, of $f$ is defined by

$$
\omega_{1}(f, \delta)=\sup _{|h| \leq \delta x, x+h \in[0, \infty)} \sup |f(x+h)-f(x)|
$$

and second modulus of smoothness of $f$ is defined by

$$
\omega_{2}(f, \delta)=\sup _{|h| \leq \delta x+2 h \in[0, \infty)} \sup |f(x+2 h)-2 f(x+h)+f(x)|
$$

We have the following well known property of the modulus of smoothness (see, e.g., [3, p. 266, Lemma 5.1.1]).

Remark 2.1. If $f \in U C_{B}[0, \infty)$, then $\lim _{\delta \rightarrow 0^{+}} \omega_{k}(f, \delta)=0$ for $k=1,2$.
For convenience, we need the following Peetre's $K$-functional defined by

$$
K(f, \delta)=\inf _{g \in C_{B}^{2}[0, \infty)}\left\{\|f-g\|+\delta\left\|g^{\prime \prime}\right\|\right\}
$$

where $\delta>0$ and

$$
C_{B}^{2}[0, \infty)=\left\{g \in C_{B}[0, \infty): g^{\prime}, g^{\prime \prime} \in C_{B}[0, \infty)\right\}
$$

Note that the modulus of smoothness and the $K$-functional of an $f \in C_{B}[0, \infty)$ are related to each other as in the following sense: There exist positive constants $C_{1}$ and $C_{2}$ such that

$$
\begin{equation*}
C_{1} \omega_{2}(f, \delta) \leq K\left(f, \delta^{2}\right) \leq C_{2} \omega_{2}(f, \delta) \tag{2.1}
\end{equation*}
$$

(see, e.g., [8, p. 177, Theorem 2.4]).
Below, we present a quantitative estimate to reach to the subsequent result concerning the rate of the approximation by $\left\{S_{n, p}^{\beta_{n}}(f ; x)\right\}_{n \geq 1}$.

Theorem 2.1. Let $p \in \mathbb{N}_{0}$ be fixed, $0 \leq \beta<1$ and $f \in C_{B}[0, \infty)$. Then, for each $x \in(0, \infty)$, one has

$$
\begin{equation*}
\left|S_{n, p}^{\beta}(f ; x)-f(x)\right| \leq \omega_{1}\left(f,\left(\beta+\frac{p}{n}\right) \frac{x}{1-\beta}\right)+C \omega_{2}\left(f, \delta_{n, p}^{\beta}(x)\right) \tag{2.2}
\end{equation*}
$$

where $C>0$ is a positive constant and

$$
\begin{equation*}
\delta_{n, p}^{\beta}(x):=\frac{1}{2} \sqrt{\left(\beta+\frac{p}{n}\right)^{2} \frac{x^{2}}{(1-\beta)^{2}}+\left(1+\frac{p}{n}\right) \frac{x}{2 n(1-\beta)^{3}}} . \tag{2.3}
\end{equation*}
$$

Proof. Consider an auxiliary operator

$$
\begin{equation*}
\bar{S}_{n, p}^{\beta}(f ; x):=S_{n, p}^{\beta}(f ; x)+f(x)-f\left(\left(1+\frac{p}{n}\right) \frac{x}{1-\beta}\right) \tag{2.4}
\end{equation*}
$$

for $f \in C_{B}[0, \infty), n \in \mathbb{N}$. In this case, $\bar{S}_{n, p}^{\beta}$ are linear and positive and each operator preserves the linear functions. Now, let $g \in C_{B}^{2}[0, \infty)$. From Taylor's formula about an arbitrary fixed point $x$, one has

$$
\begin{equation*}
g(t)=g(x)+g^{\prime}(x)(t-x)+\int_{x}^{t}(t-y) g^{\prime \prime}(y) d y \tag{2.5}
\end{equation*}
$$

for $t \in[0, \infty)$. Application of the operators $\bar{S}_{n, p}^{\beta}$ on both sides of (2.5) gives that

$$
\begin{equation*}
\bar{S}_{n, p}^{\beta}(g ; x)-g(x)=g^{\prime}(x) \bar{S}_{n, p}^{\beta}(t-x ; x)+\bar{S}_{n, p}^{\beta}\left(\int_{x}^{t}(t-y) g^{\prime \prime}(y) d y ; x\right) \tag{2.6}
\end{equation*}
$$

Taking (2.4) into account for $f(t)=\int_{x}^{t}(t-y) g^{\prime \prime}(y) d y$, expression (2.6) reduces to $\bar{S}_{n, p}^{\beta}(g ; x)-g(x)=S_{n, p}^{\beta}\left(\int_{x}^{t}(t-y) g^{\prime \prime}(y) d y ; x\right)-\int_{x}^{\left(1+\frac{p}{n}\right)}\left[\left(1+\frac{x}{n}\right) \frac{x}{1-\beta}-y\right] g^{\prime \prime}(y) d y$. Using the fact

$$
\left|\int_{x}^{\left(1+\frac{p}{n} \frac{x}{1-\beta}\right.}\left[\left(1+\frac{p}{n}\right) \frac{x}{1-\beta}-y\right] g^{\prime \prime}(y) d y\right| \leq \frac{1}{2}\left(S_{n, p}^{\beta}\left(\varphi_{x}^{1} ; x\right)\right)^{2}\left\|g^{\prime \prime}\right\|
$$

by Lemma 1.3, we obtain

$$
\begin{aligned}
& \left|\bar{S}_{n, p}^{\beta}(g ; x)-g(x)\right| \\
\leq & S_{n, p}^{\beta}\left(\left|\int_{x}^{t}(t-y) g^{\prime \prime}(y) d y\right| ; x\right)+\left|\int_{x}^{\left(1+\frac{p}{n}\right) \frac{x}{1-\beta}}\left[\left(1+\frac{p}{n}\right) \frac{x}{1-\beta}-y\right]\right| g^{\prime \prime}(y)|d y| \\
\leq & \frac{\left\|g^{\prime \prime}\right\|}{2}\left[S_{n, p}^{\beta}\left(\varphi_{x}^{2} ; x\right)+\left(S_{n, p}^{\beta}\left(\varphi_{x}^{1} ; x\right)\right)^{2}\right] \\
(2 \Rightarrow) & \frac{\left\|g^{\prime \prime}\right\|}{2}\left[2\left(\beta+\frac{p}{n}\right)^{2} \frac{x^{2}}{(1-\beta)^{2}}+\left(1+\frac{p}{n}\right) \frac{x}{n(1-\beta)^{3}}\right] .
\end{aligned}
$$

On the other hand, from (2.4) and Lemma 1.2, it can be easily obtained that

$$
\begin{equation*}
\left|\bar{S}_{n, p}^{\beta}(f ; x)\right| \leq\left|S_{n, p}^{\beta}(f ; x)\right|+2\|f\| \leq 3\|f\| \tag{2.8}
\end{equation*}
$$

for $f \in C_{B}[0, \infty)$. Thus, taking (2.4), (2.7) and (2.8) into account, for $f, g \in$ $C_{B}[0, \infty)$ one has

$$
\begin{aligned}
& \left|S_{n, p}^{\beta}(f ; x)-f(x)\right| \\
\leq & \left|\bar{S}_{n, p}^{\beta}(f-g ; x)-(f-g)(x)\right|+\left|\bar{S}_{n, p}^{\beta}(g ; x)-g(x)\right| \\
& +\left|f(x)-f\left(\left(1+\frac{p}{n}\right) \frac{x}{1-\beta}\right)\right| \\
\leq & \omega_{1}\left(f,\left(\beta+\frac{p}{n}\right) \frac{x}{1-\beta}\right) \\
& +4\left\{\|f-g\|+\frac{1}{4}\left[\left(\beta+\frac{p}{n}\right)^{2} \frac{x^{2}}{(1-\beta)^{2}}+\left(1+\frac{p}{n}\right) \frac{x}{2 n(1-\beta)^{3}}\right]\left\|g^{\prime \prime}\right\|\right\} .
\end{aligned}
$$

Finally, taking infimum over all $g \in C_{B}^{2}[0, \infty)$ on the right hand-side of the last inequality and applying (2.1), we get

$$
\begin{aligned}
\left|S_{n, p}^{\beta}(f ; x)-f(x)\right| & \leq \omega_{1}\left(f,\left(\beta+\frac{p}{n}\right) \frac{x}{1-\beta}\right)+K\left(f,\left(\delta_{n, p}^{\beta}(x)\right)^{2}\right) \\
& \leq \omega_{1}\left(f,\left(\beta+\frac{p}{n}\right) \frac{x}{1-\beta}\right)+C \omega_{2}\left(f, \delta_{n, p}^{\beta}(x)\right)
\end{aligned}
$$

where $\delta_{n, p}^{\beta}(x)$ is given by (2.3).
Note that the case $p=0$ in the above theorem reduces to Theorem 2 in [2].
Taking Remark 2.1 and (2.2) into account, we reach to the following conclusion:
Corollary 2.1. i) If $\beta$ is taken as a sequence $\beta_{n}$ such that $0 \leq \beta_{n}<1$ and $\lim _{n \rightarrow \infty} \beta_{n}=0$ and $f \in U C_{B}[0, \infty)$, then one gets $\lim _{n \rightarrow \infty} S_{n, p}^{\beta_{n}}(f ; x)=f(x)$ on $[0, \infty)$ and the order of the approximation does not exceed to that of $\omega_{1}\left(f,\left(\beta_{n}+\frac{p}{n}\right) \frac{x}{1-\beta_{n}}\right)+$ $C \omega_{2}\left(f, \delta_{n, p}^{\beta_{n}}(x)\right)$.
ii) If $\beta$ is taken as a sequence $\beta_{n}$ such that $0 \leq \beta_{n}<1$ and $\lim _{n \rightarrow \infty} \beta_{n}=0$ and $f \in C_{B}[0, \infty)$, then $\left\{S_{n, p}^{\beta_{n}}(f)\right\}_{n \geq 1}$ converges uniformly to $f$ on $[a, b], 0 \leq a<b<$ $\infty$, by the well known Korovkin theorem.

## 3. A Voronovskaja-type result

In [9], Farcaş obtained the following Voronovskaja-type result for the Jain operator $P_{n}^{[\beta]}$ given by (1.2):

$$
\lim _{n \rightarrow \infty} n\left\{P_{n}^{\left[\beta_{n}\right]}(f ; x)-f(x)\right\}=\frac{x}{2} f^{\prime \prime}(x), \quad x>0
$$

for $f \in C_{2}[0, \infty)$, the space of all continuous functions having continuous second order derivative, where $0 \leq \beta_{n}<1$ is a sequence such that $\lim _{n \rightarrow \infty} \beta_{n}=0$.

Note that a Voronovskaja-type result for a generalization of the Jain operators was obtained by Olgun et al. [18]. On the other hand, a Voronovskaja-type theorem as well as its a generalized form for Schurer setting of the Szász-Mirakjan operators were obtained by Sikkema in [20, p. 333].

In this part, we investigate a Voronovskaja-type result for the Jain-Schurer operators $S_{n, p}^{\beta}, n \in \mathbb{N}$.

Theorem 3.1. Let $p \in \mathbb{N}_{0}$ be fixed and $0 \leq \beta_{n}<1$ be a sequence such that $\lim _{n \rightarrow \infty} n \beta_{n}=0$. If $f$ is bounded and continuous on $[0, \infty)$ and has the second order derivative at some $x \in(0, \infty)$, then one has

$$
\lim _{n \rightarrow \infty} n\left\{S_{n, p}^{\beta_{n}}(f ; x)-f(x)\right\}=p x f^{\prime}(x)+\frac{x}{2} f^{\prime \prime}(x)
$$

Proof. From Taylor's formula, one has

$$
\begin{equation*}
f(t)=f(x)+f^{\prime}(x)(t-x)+\frac{1}{2} f^{\prime \prime}(x)(t-x)^{2}+h(t-x)(t-x)^{2} \tag{3.1}
\end{equation*}
$$

at the fixed point $x \in[0, \infty)$, where $h(t-x)$ is bounded for all $t \in[0, \infty)$ and $\lim _{t \rightarrow x} h(t-x)=0$. Application of the operators $S_{n, p}^{\beta}$ to (3.1) implies

$$
\begin{aligned}
n\left[S_{n, p}^{\beta_{n}}(f ; x)-f(x)\right]= & f^{\prime}(x) n S_{n, p}^{\beta_{n}}(t-x ; x)+\frac{1}{2} f^{\prime \prime}(x) n S_{n, p}^{\beta_{n}}\left((t-x)^{2} ; x\right) \\
& +n S_{n, p}^{\beta_{n}}\left(h(t-x)(t-x)^{2} ; x\right)
\end{aligned}
$$

Using the facts $\lim _{n \rightarrow \infty} n \beta_{n}=0$ and Lemma 1.3, it readily follows that

$$
\lim _{n \rightarrow \infty} n S_{n, p}^{\beta_{n}}(t-x ; x)=p x
$$

and

$$
\lim _{n \rightarrow \infty} n S_{n, p}^{\beta_{n}}\left((t-x)^{2} ; x\right)=x
$$

Hence, we have
$\lim _{n \rightarrow \infty} n\left(S_{n, p}^{\beta_{n}}(f ; x)-f(x)\right)=p x f^{\prime}(x)+\frac{x}{2} f^{\prime \prime}(x)+\lim _{n \rightarrow \infty} n S_{n, p}^{\beta_{n}}\left(h(t-x)(t-x)^{2} ; x\right)$.
It suffices to prove that $\lim _{n \rightarrow \infty} n S_{n, p}^{\beta_{n}}\left(h(t-x)(t-x)^{2} ; x\right)=0$. Indeed, defining $h(0)=0$ and taking the fact $\lim _{t \rightarrow x} h(t-x)=0$ into account, we get that $h$ is continuous at $x$. Hence, for each $\varepsilon>0$, there is a $\delta>0$ such that $|h(t-x)|<$ $\varepsilon$ for all $t$ satisfying $|t-x|<\delta$. On the other hand, since $h(t-x)$ is bounded on $[0, \infty)$, there is an $M>0$ such that $|h(t-x)| \leq M$ for all $t$. Therefore, we may write $|h(t-x)| \leq M \frac{(t-x)^{2}}{\delta^{2}}$ when $|t-x| \geq \delta$. So, these arguments enable one to write $|h(t-x)| \leq \varepsilon+M \frac{(t-x)^{2}}{\delta^{2}}$ for all $t$. The monotonicity and linearity of $S_{n, p}^{\beta_{n}}$ give that

$$
\begin{aligned}
S_{n, p}^{\beta_{n}}\left(h(t-x)(t-x)^{2} ; x\right) & \leq \varepsilon S_{n, p}^{\beta_{n}}\left((t-x)^{2} ; x\right)+\frac{M}{\delta^{2}} S_{n, p}^{\beta_{n}}\left((t-x)^{4} ; x\right) \\
& =\varepsilon S_{n, p}^{\beta_{n}}\left(\varphi_{x}^{2} ; x\right)+\frac{M}{\delta^{2}} S_{n, p}^{\beta_{n}}\left(\varphi_{x}^{4} ; x\right)
\end{aligned}
$$

Making use of Lemma 1.3, with $\beta=\beta_{n}$,

$$
\lim _{n \rightarrow \infty} n S_{n, p}^{\beta_{n}}\left(h(t-x)(t-x)^{2} ; x\right)=0
$$

by the hypothesis on $\beta_{n}$, which completes the proof.

## 4. A Retaining Property

Recall that $A$ continuous and non-negative function $\omega$ defined on $[0, \infty)$ is called $a$ modulus of continuity, if each of the following conditions is satisfied:
i) $\omega(u+v) \leq \omega(u)+\omega(v)$ for $u, v, u+v \in[0, \infty)$, i.e., $\omega$ is semi-additive,
ii) $\omega(u) \geq \omega(v)$ for $u \geq v>0$, i.e., $\omega$ is non-decreasing,
iii) $\lim _{u \rightarrow 0^{+}} \omega(u)=\omega(0)=0,([15, \mathrm{p} .106])$.

In [13], Li proved that each Bernstein polynomial preserves the properties of modulus of continuity on $[0,1]$. Motivated by this result, in this section we will show that each Jain-Schurer operator has this preservation property as well. In the proof, we need the following Jensen formula

$$
\begin{equation*}
(u+v)(u+v+m \beta)^{m-1}=\sum_{k=0}^{m}\binom{m}{k} u(u+k \beta)^{k-1} v[v+(m-k) \beta]^{m-k-1} \tag{4.1}
\end{equation*}
$$

where $u, v$, and $\beta \in \mathbb{R}$ (see, e.g., [3, p. 326]).

Theorem 4.1. Let $p \in \mathbb{N}_{0}$ be fixed and $0 \leq \beta<1$. If $\omega$ is a bounded modulus of continuity on $[0, \infty)$, then for each $n \in \mathbb{N}, S_{n, p}^{\beta}(\omega ; x)$ is also a modulus of continuity.

Proof. Let $x, y \in[0, \infty)$ and $x \leq y$. From the definition of $S_{n}^{\beta}$, we have

$$
\begin{equation*}
S_{n, p}^{\beta}(\omega ; y)=\sum_{j=0}^{\infty} \omega\left(\frac{j}{n}\right) \frac{(n+p) y[(n+p) y+j \beta]^{j-1}}{j!} e^{-[(n+p) y+j \beta]} \tag{4.2}
\end{equation*}
$$

Taking $u=(n+p) x, v=(n+p) y-(n+p) x$ and $m=j$ in (4.1), we obtain

$$
\begin{aligned}
& (n+p) y((n+p) y+j \beta)^{j-1} \\
= & \sum_{k=0}^{j}\binom{j}{k}(n+p) x[(n+p) x+k \beta]^{k-1} \\
& \times(n+p)(y-x)[(n+p)(y-x)+(j-k) \beta]^{j-k-1} .
\end{aligned}
$$

Substituting this expression into (4.2) we get

$$
\begin{align*}
S_{n, p}^{\beta}(\omega ; y)= & \sum_{j=0}^{\infty} \sum_{k=0}^{j} \omega\left(\frac{j}{n}\right)\binom{j}{k} \frac{1}{j!}(n+p) x[(n+p) x+k \beta]^{k-1} \\
& \times(n+p)(y-x)[(n+p)(y-x)+(j-k) \beta]^{j-k-1} e^{-[(n+p) y+j \beta]} \\
= & \sum_{k=0}^{\infty} \sum_{j=k}^{\infty} \omega\left(\frac{j}{n}\right) \frac{1}{k!(j-k)!}(n+p) x[(n+p) x+k \beta]^{k-1} \\
& \times(n+p)(y-x)[(n+p)(y-x)+(j-k) \beta]^{j-k-1} e^{-[(n+p) y+j \beta]} \\
= & \sum_{k=0}^{\infty} \sum_{l=0}^{\infty} \omega\left(\frac{k+l}{n}\right) \frac{1}{k!l!}(n+p) x[(n+p) x+k \beta]^{k-1} \\
& \times(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1} e^{-[(n+p) y+k \beta+l \beta]} . \tag{4.3}
\end{align*}
$$

On the other hand, from (1.3), we have

$$
e^{(n+p)(y-x)}=\sum_{l=0}^{\infty} \frac{(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1}}{l!} e^{-l \beta}
$$

Therefore, $S_{n, p}^{\beta}(\omega ; x)$ may be written as

$$
\begin{align*}
S_{n, p}^{\beta}(\omega ; x)= & \sum_{k=0}^{\infty} \omega\left(\frac{k}{n}\right) \frac{(n+p) x[(n+p) x+k \beta]^{k-1}}{k!} e^{-[(n+p) x+k \beta]} \\
= & \sum_{k=0}^{\infty} \omega\left(\frac{k}{n}\right) \frac{(n+p) x[(n+p) x+k \beta]^{k-1}}{k!} e^{-[(n+p) y+k \beta]} e^{(n+p)(y-x)} \\
= & \sum_{k=0}^{\infty} \sum_{l=0}^{\infty} \omega\left(\frac{k}{n}\right) \frac{1}{k!l!}(n+p) x[(n+p) x+k \beta]^{k-1} \\
& \times(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1} e^{-[(n+p) y+k \beta+l \beta]} . \tag{4.4}
\end{align*}
$$

Subtracting (4.4) from (4.3), we obtain

$$
\begin{align*}
& S_{n, p}^{\beta}(\omega ; y)-S_{n, p}^{\beta}(\omega ; x) \\
= & \sum_{k=0}^{\infty} \sum_{l=0}^{\infty}\left\{\omega\left(\frac{k+l}{n}\right)-\omega\left(\frac{k}{n}\right)\right\} \frac{1}{k!l!}(n+p) x[(n+p) x+k \beta]^{k-1} \\
& \times(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1} e^{-[(n+p) y+(k+l) \beta]} . \tag{4.5}
\end{align*}
$$

Using the semi-additivity property of $\omega$, we get

$$
\begin{aligned}
& S_{n, p}^{\beta}(\omega ; y)-S_{n, p}^{\beta}(\omega ; x) \\
\leq & \sum_{k=0}^{\infty} \frac{(n+p) x[(n+p) x+k \beta]^{k-1}}{k!} e^{-k \beta} \\
& \times \sum_{l=0}^{\infty} \omega\left(\frac{l}{n}\right) \frac{(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1}}{l!} e^{-[(n+p) y+l \beta]} \\
= & e^{(n+p) x} \sum_{l=0}^{\infty} \omega\left(\frac{l}{n}\right) \frac{(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1}}{l!} e^{-[(n+p) y+l \beta]} \\
= & \sum_{l=0}^{\infty} \omega\left(\frac{l}{n}\right) \frac{(n+p)(y-x)[(n+p)(y-x)+l \beta]^{l-1}}{l!} e^{-[(n+p)(y-x)+l \beta]} \\
= & S_{n, p}^{\beta}(\omega ; y-x),
\end{aligned}
$$

which shows the semi-additivity of $S_{n, p}^{\beta}$. From (4.5) it readily follows that $S_{n, p}^{\beta}(\omega ; y) \geq$ $S_{n, p}^{\beta}(\omega ; x)$ for $y \geq x$, i.e., $S_{n, p}^{\beta}$ is non-decreasing. Moreover, since the series is uniformly convergent, it follows that $\lim _{x \rightarrow 0^{+}} S_{n, p}^{\beta}(\omega ; x)=S_{n, p}^{\beta}(\omega ; 0)=\omega(0)=0$. This comples the proof.

## 5. Monotonicity of the sequence of the Jain-Schurer operators

In [5], Cheney and Sharma proved that the sequence of Szász-Mirakjan operators $P_{n}^{[0]}(f)$ is non-increasing in $n$, when $f$ is convex. The purpose of this section is to observe the monotonicity of the sequence of the Jain-Schurer operators when the attached function is convex and non-decreasing and $p \neq 0$. In the case $p=0$, we obtain monotonicity of the sequence of Jain operators in $n$ when $f$ is convex. For the proof, we further need the following Abel-Jensen formula

$$
\begin{equation*}
(u+v+m \beta)^{m}=\sum_{k=0}^{m}\binom{m}{k}(u+k \beta)^{k} v[v+(m-k) \beta]^{m-k-1} \tag{5.1}
\end{equation*}
$$

for non-negative real number $\beta$, where $u, v \in \mathbb{R}$ and $m \geq 1$ (see, e.g., [21]). Reasoning as in [5], we present the following result:

Theorem 5.1. Let $f$ be a non-decreasing and convex function on $[0, \infty)$. Then, for all $n, S_{n, p}^{\beta}(f)$ is non-increasing in $n$ when $p \neq 0$. For the case $p=0$, the same result holds when $f$ is only convex on $[0, \infty)$.

Proof. From (1.3), with $\alpha=x$, it is obvious that

$$
\begin{equation*}
e^{x}=\sum_{k=0}^{\infty} \frac{x(x+k \beta)^{k-1}}{k!} e^{-k \beta} \tag{5.2}
\end{equation*}
$$

Since $S_{n, p}^{\beta}(f ; 0)=f(0)$, we study only for $x>0$. Taking the definition of $S_{n, p}^{\beta}$ and (5.2) into consideration, one has

$$
\begin{aligned}
& S_{n, p}^{\beta}(f ; x)-S_{n+1, p}^{\beta}(f ; x) \\
= & e^{x} \sum_{k=0}^{\infty} f\left(\frac{k}{n}\right) \frac{(n+p) x[(n+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]} \\
& -\sum_{k=0}^{\infty} f\left(\frac{k}{n+1}\right) \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]} \\
= & \sum_{l=0}^{\infty} \frac{x(x+l \beta)^{l-1}}{l!} e^{-l \beta} \sum_{k=0}^{\infty} f\left(\frac{k}{n}\right) \frac{(n+p) x[(n+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]} \\
& -\sum_{k=0}^{\infty} f\left(\frac{k}{n+1}\right) \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]} .
\end{aligned}
$$

By simple calculations, one can write

$$
\begin{aligned}
& \left(5.3 \Psi_{n, p}^{\beta}(f ; x)-S_{n+1, p}^{\beta}(f ; x)\right. \\
= & \sum_{l=0}^{\infty} \frac{x(x+l \beta)^{l-1}}{l!} e^{-l \beta} \sum_{k=l}^{\infty} f\left(\frac{k-l}{n}\right) \frac{(n+p) x[(n+p) x+(k-l) \beta]^{k-l-1}}{(k-l)!} e^{-[(n+1+p) x+(k-l) \beta]} \\
& -\sum_{k=0}^{\infty} f\left(\frac{k}{n+1}\right) \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]} \\
= & \sum_{k=0}^{\infty} e^{-[(n+1+p) x+k \beta]}\left\{\sum_{l=0}^{k} f\left(\frac{k-l}{n}\right) \frac{(n+p) x[(n+p) x+(k-l) \beta]^{k-l-1}}{(k-l)!} \frac{x(x+l \beta)^{l-1}}{l!}\right. \\
& \left.-f\left(\frac{k}{n+1}\right) \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!}\right\} \\
= & \sum_{k=0}^{\infty} e^{-[(n+1+p) x+k \beta]}\left\{\sum_{l=0}^{k} f\left(\frac{l}{n}\right) \frac{(n+p) x[(n+p) x+l \beta]^{l-1}}{l!} \frac{x[x+(k-l) \beta]^{k-l-1}}{(k-l)!}\right. \\
& \left.-f\left(\frac{k}{n+1}\right) \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!}\right\} \\
= & \sum_{k=0}^{\infty} \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]} \times \\
& \left\{\sum_{l=0}^{k}\binom{k}{l} \frac{(n+p) x[(n+p) x+l \beta]^{l-1} x[x+(k-l) \beta]^{k-l-1}}{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}} f\left(\frac{l}{n}\right)-f\left(\frac{k}{n+1}\right)\right\} .
\end{aligned}
$$

Now, it only remains to show that the curly bracket in the last formula must be non-negative. For this, we denote

$$
\alpha_{l}:=\binom{k}{l} \frac{(n+p) x[(n+p) x+l \beta]^{l-1} x[x+(k-l) \beta]^{k-l-1}}{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}>0,
$$

and

$$
x_{l}=\frac{l}{n}
$$

for $l=0,1, \ldots, k$. Now, replacing $u$ with $(n+p) x, v$ with $x, m$ with $k$ and $k$ with $l$ in (4.1) we evidently get
(5.4)

$$
\sum_{l=0}^{k} \alpha_{l}=\frac{1}{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}(n+1+p) x[(n+1+p) x+k \beta]^{k-1}=1
$$

On the other hand, it follows that

$$
\begin{align*}
\sum_{l=0}^{k} \alpha_{l} x_{l}= & \frac{1}{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}} \times \\
& \sum_{l=0}^{k}\binom{k}{l} \frac{l}{n}(n+p) x[(n+p) x+l \beta]^{l-1} x[x+(k-l) \beta]^{k-l-1} \\
= & \frac{k(n+p) x}{n(n+1+p) x[(n+1+p) x+k \beta]^{k-1}} \times \\
& \sum_{l=0}^{k-1}\binom{k-1}{l}[(n+p) x+\beta+l \beta]^{l} x[x+(k-l-1) \beta]^{k-l-2} . \tag{5.5}
\end{align*}
$$

Making use of the Abel-Jensen formula given by (5.1) for $u=(n+p) x+\beta, v=$ $x, k=l, m=k-1,(5.5)$ reduces to

$$
\begin{align*}
\sum_{l=0}^{k} \alpha_{l} x_{l} & =\frac{k(n+p) x}{n(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}[(n+1+p) x+k \beta]^{k-1} \\
& =\frac{k(n+p)}{n(n+1+p)} \tag{5.6}
\end{align*}
$$

Taking into account (5.4), (5.6) and the convexity of $f$, (5.3) reduces to

$$
\begin{aligned}
& \text { (5.7) } S_{n, p}^{\beta}(f ; x)-S_{n+1, p}^{\beta}(f ; x) \\
& =\sum_{k=0}^{\infty} \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]}\left\{\sum_{l=0}^{k} \alpha_{l} f\left(\frac{l}{n}\right)-f\left(\frac{k}{n+1}\right)\right\} \\
& \geq \sum_{k=0}^{\infty} \frac{(n+1+p) x[(n+1+p) x+k \beta]^{k-1}}{k!} e^{-[(n+1+p) x+k \beta]}\left\{f\left(\frac{k(n+p)}{n(n+1+p)}\right)-f\left(\frac{k}{n+1}\right)\right\} .
\end{aligned}
$$

It is obvious that when $p=0,(5.7)$ gives the non-negativity of $S_{n, 0}^{\beta}(f ; x)-$ $S_{n+1,0}^{\beta}(f ; x)$ under the convexity of $f$, which means that the sequence of Jain operators is non-increasing in $n$ under the convexity of the function. On the other hand, for $p \in \mathbb{N}$ it follows that

$$
\frac{k(n+p)}{n(n+1+p)}=\frac{k}{n+1} \frac{n+1}{n} \frac{n+p}{n+1+p}=\frac{k}{n+1} \frac{1+\frac{p}{n}}{\left(1+\frac{p}{n+1}\right)}
$$

Hence, one has

$$
\frac{k(n+p)}{n(n+1+p)} \geq \frac{k}{n+1}
$$

by the fact that $\frac{1+\frac{p}{n}}{1+\frac{n}{n+1}} \geq 1$. Then, the result follows directly from the non-decreasingness of $f$.

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Nursel Çetin
Turkish State Meteorological Service
Research Department
06560, Ankara, Turkey
nurselcetin07@gmail.com

Gülen Başcanbaz-Tunca
Ankara University
Faculty of Science
Department of Mathematics 06100, Tandogan Ankara, Turkey

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tunca@science.ankara.edu.tr
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# AUTHOMATED METHOD FOR DESIGNING FUZZY SYSTEMS * 

Ivana Micić, Nada Damljanović and Zorana Jančić

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#### Abstract

The paper presents a method for building fuzzy systems using the inputoutput data that can be obtained from the examples. Using this method, a rulebased system is created, where fuzzy logic depends on the opinions and preferences of decision-makers involved in the process. Some advantages of the proposed method are highlighted. We have provided a practical example to illustrate the application of the process.


Keywords: fuzzy systems; rule-based system; fuzzy rule; membership function.

## 1. Introduction

Zadeh's fuzzy rule based systems deal with fuzzy rules instead of classical logic rules. Nowadays, they have been successfully used for modeling and control in different fields and industries $[1,2,5,15,16]$.

Fuzzy rule based systems with fuzzifier and defuzzifier introduced by Mamdani $[9,10]$ are commonly known as fuzzy logic controllers. Mamdani fuzzy rule based systems deal with real-valued inputs and outputs, and therefore, they can be used in a wide range of real-world applications. The behavior of the system is guided by linguistic rules with the "IF-THEN" form whose premises and consequents are composed of fuzzy logic statements [3, 12, 14]. More on linguistic Mamdani-type fuzzy rule-based systems can be found in [17].

One of drawbacks of Mamdani fuzzy rule based systems can be viewed in a fact that good performance on input-output training data do not nonsensically led to good performance on novel inputs $[4,6,11]$. Therefore, a construction of fuzzy functions and corresponding base of rules based on inclusion of expert knowledge into the process is proposed.

The model presented in this paper is shown to be very good, because of its flexibility, therefore it can be very easy for implementing and application in various

[^4]fields. In this work, trough one illustrative example, it will be shown that greater number of functions for presenting the input data will give much better results, and therefore, the flexibility od the model is limited by the lower bound in the number of functions for presenting input data.

## 2. Automated method for designing fuzzy systems based on learning from example

The input of the observed fuzzy system is a set of $N$ input-output pairs, of the form

$$
\begin{equation*}
\left\{\left(X_{0}^{p}, y_{0}^{p}\right)\right\}, \quad p \in 1, \ldots, N \tag{2.1}
\end{equation*}
$$

where $X_{0}^{p} \in U=\left[\alpha_{1}, \beta_{1}\right] \times \cdots \times\left[\alpha_{n}, \beta_{n}\right] \subset R^{n}$ and $y_{0}^{p} \in V=\left[\alpha_{y}, \beta_{y}\right] \subset R$. Clearly, the input of a fuzzy system (2.1) is the collection of data given by Table 2.1.

Table 2.1: Input-output data collection

| Input | $C_{1}$ | $C_{2}$ | $\ldots$ | $C_{i}$ | $\ldots$ | $C_{n}$ | Output |
| :---: | :---: | :---: | :--- | :---: | :--- | :--- | :---: |
| $X_{0}^{1}$ | $x_{01}^{1}$ | $x_{02}^{1}$ | $\ldots$ | $x_{0 i}^{1}$ | $\ldots$ | $x_{0 n}^{1}$ | $y_{0}^{1}$ |
| $X_{0}^{2}$ | $x_{01}^{2}$ | $x_{02}^{2}$ | $\ldots$ | $x_{0 i}^{2}$ | $\ldots$ | $x_{0 n}^{2}$ | $y_{0}^{2}$ |
|  |  |  | $\ldots$ |  | $\ldots$ |  |  |
| $X_{0}^{p}$ | $x_{01}^{p}$ | $x_{02}^{p}$ | $\ldots$ | $x_{0 i}^{p}$ | $\ldots$ | $x_{0 n}^{p}$ | $y_{0}^{p}$ |
|  |  |  | $\ldots$ |  | $\ldots$ |  |  |
| $X_{0}^{N}$ | $x_{01}^{N}$ | $x_{02}^{N}$ | $\ldots$ | $x_{0 i}^{N}$ | $\ldots$ | $x_{0 n}^{N}$ | $y_{0}^{N}$ |

Designing the fuzzy system based on these input-output data collection can be described in the the following five steps.

## Step 1. Experts opinion

For each $i \in\{1,2, \ldots, n\}$ and corresponding attribute $C_{i}$, values represented in the $i$ th column of Table 2.1 can have different importance to a decision expert. Some values are extremely important, while others are totally unacceptable. On the other hand, different decision experts can have different intuition and preferences on what's important. Therefore, $N_{i}$ decision experts are involved to express their preference on attribute $C_{i}$.

For each $i \in\{1,2, \ldots, n\}$ and $j \in\left\{1,2, \ldots, N_{i}\right\}$, the $j$ th expert on attribute $C_{i}$ choose four elements $a_{i}^{j}, b_{i}^{j}, c_{i}^{j}, d_{i}^{j} \in\left[\alpha_{i}, \beta_{i}\right]$.

Table 2.2: Expert's preference on atributes

$$
\begin{array}{|l|l|l|l|}
\hline a_{i}^{j} & b_{i}^{j} & c_{i}^{j} & d_{i}^{j} \\
\hline
\end{array}
$$

For example, let attribute $C_{i}$ represent price of some article (or service) which can generally range between $\alpha_{i}$ and $\beta_{i}$. Given values $a_{i}^{j}, b_{i}^{j}, c_{i}^{j}$ and $d_{i}^{j}$ have the
following meaning: If the price is lower than $a_{i}$ or it is higher than $d_{i}$, then we are not interested in buying that article (too cheap or too expensive items are not interesting to us). If the price is between $b_{i}^{j}$ and $c_{i}^{j}$, then we are absolutely interested in buying the article (shopping surely). As price goes from $a_{i}$ to $b_{i}$, we are increasingly interested for buying it, and if price goes from $c_{i}$ to $d_{i}$ our interest in the purchase of item drops.

In this way, for each attribute $C_{i}(i=1,2, \ldots, n)$, we have determined $N_{i}$ fuzzy sets

$$
\begin{equation*}
A_{i}^{j}:\left[\alpha_{i}, \beta_{i}\right] \rightarrow[0,1], \quad j=1,2, \ldots, N_{i} \tag{2.2}
\end{equation*}
$$

as follows:

$$
A_{i}^{j}(x)=\left\{\begin{array}{cc}
0, & \alpha_{i} \leqslant x \leq a_{i}^{j} \text { or } d_{i}^{j} \leqslant x \leq \beta_{i} ; \\
\frac{x-a_{i}^{j}}{b_{i}^{j}-a_{i}^{j}}, & a_{i}^{j} \leqslant x \leq b_{i}^{j} ; \\
1, & b_{i}^{j} \leqslant x \leq c_{i}^{j} ; \\
\frac{x-d_{i}^{j}}{c_{i}^{j}-d_{i}^{j}}, & c_{i}^{j} \leqslant x \leq d_{i}^{j} .
\end{array}\right.
$$

It is assumed that, for each $i=1,2, \ldots, n$, the set of fuzzy functions (2.2) is complete in $\left[\alpha_{i}, \beta_{i}\right]$, i.e., for every $x_{i} \in\left[\alpha_{i}, \beta_{i}\right]$, there exists $A_{i}^{j}$ such that $\mu_{A_{i}^{j}}\left(x_{i}\right) \neq 0$.

With similar arguments, $N_{y}$ decision experts are involved to express their preference on output column $\left(y_{0}^{1}, y_{0}^{2}, \ldots, y_{0}^{N}\right)^{T}$.

Table 2.3: Expert's preference on output

| $a_{i}$ | $b_{i}$ | $c_{i}$ | $d_{i}$ |
| :--- | :--- | :--- | :--- |

Consequently, $N_{y}$ fuzzy sets

$$
\begin{equation*}
B^{j}:\left[\alpha_{y}, \beta_{y}\right] \rightarrow[0,1], \quad j=1,2, \ldots, N_{y} \tag{2.3}
\end{equation*}
$$

are defined in the following way:

$$
B^{j}(x)=\left\{\begin{array}{cc}
0, & \alpha_{y} \leqslant x \leq a^{j} \text { or } d^{j} \leqslant x \leqslant \beta_{y} \\
\frac{x-a^{j}}{b^{j}-a^{j}}, & a^{j} \leqslant x \leq b^{j} ; \\
1, & b^{j} \leqslant x \leq c^{j} ; \\
\frac{x-d^{j}}{c^{j}-d^{j}}, & c^{j} \leqslant x \leq d^{j}
\end{array}\right.
$$

Again, the assumption is that they are complete in $\left[\alpha_{y}, \beta_{y}\right]$.
One can notice that in the case of incompleteness of obtained fuzzy sets (2.2) or (2.3), the number of experts being examined must increase. Also, let us notice that
obtained fuzzy sets are trapezoidal, and in a naturally way they can be transformed to triangular fuzzy sets or singletons.

Step 2. Rules generated by input-output data
In this step, for every input-output pair

$$
\left(X_{0}^{p}, y_{0}^{p}\right), \quad p=1,2, \ldots, N
$$

and corresponding inputs and output

$$
x_{0 i}^{p}, \quad i=1,2, \ldots, n \quad \text { and } \quad y_{0}^{p},
$$

we will determine the membership values

$$
A_{i}^{j}\left(x_{0 i}^{p}\right), \quad j=1,2, \ldots, N_{i}
$$

and the membership values

$$
B^{l}\left(y_{0}^{p}\right), \quad l=1,2, \ldots, N_{y}
$$

Then for every input variable $x_{0 i}^{p}, i=1,2, \ldots, n$, we will determine the fuzzy set in which $x_{0 i}^{p}$ has the largest membership value, that is, we will determine $A_{i}^{j *}$ such that

$$
A_{i}^{j *}\left(x_{0 i}^{p}\right) \geqslant A_{i}^{j}\left(x_{0 i}^{p}\right), \quad j=1,2, \ldots, N_{i}
$$

Similarly, we will determine $B^{l *}$ such that

$$
B^{l *}\left(y_{0}^{p}\right) \geqslant B^{l}\left(y_{0}^{p}\right), \quad 1=1,2, \ldots, N_{y} .
$$

Finally, we obtain a fuzzy IF-THEN rule as

$$
\begin{equation*}
\text { IF } \quad x_{1}=A_{1}^{j *} \quad \text { and } \cdots \quad \text { and } \quad x_{n}=A_{n}^{j *} \quad \text { THEN } \quad y=B^{l *} . \tag{2.4}
\end{equation*}
$$

## Step 3. Degrees of fuzzy rules

Since the number of input-output pairs is usually large, and for every pair one rule is generated, it is highly likely that there are conflicting rules, i.e., there are rules with the same IF part and different THEN part. In order to overcome this conflict, the degree to each rule generated in Step 2 is assigned and only one rule from a conflicting group that has the maximum degree is chosen. That procedure resolves the conflict problem, but also reduced the number of rules.

The degree of the rule, denoted by $D$, is defined as follows: Let the rule (2.4) be generated by a pair $\left(X_{0}^{p}, y_{0}^{p}\right)$, then its degree is defined by:

$$
\begin{equation*}
D(\text { rule })=\prod_{i=1}^{n} A_{1}^{j *}\left(x_{0 i}^{p}\right) \cdot B^{l *}\left(y_{0}^{p}\right) \tag{2.5}
\end{equation*}
$$

If the input-output pairs have different reliability and we can determine a number to asses it, we may incorporate this information into the degrees of the rules.

Specifically, suppose the input-output pair $\left(X_{0}^{p}, y_{0}^{p}\right)$ has the degree $\mu^{p} \in[0,1]$, then the degree of the rule generated by a pair $\left(X_{0}^{p}, y_{0}^{p}\right)$ is defined by:

$$
\begin{equation*}
D(\text { rule })=\prod_{i=1}^{n} A_{1}^{j *}\left(x_{0 i}^{p}\right) \cdot B^{l *}\left(y_{0}^{p}\right) \cdot \mu^{p} . \tag{2.6}
\end{equation*}
$$

In practice, an expert may check the data (if the number of input-output pairs is small) and estimate the degree $\mu^{p}$. If we cannot tell the difference among the input-output pairs, we simply choose all $\mu^{p}$ value 1 , in that way (2.6) is reduced to (2.5).

Step 4. Fuzzy rule base
The fuzzy rule base consists of the following set of rules:

1. The rules generated in Step 2 that do not conflict with any other rules;
2. The rule from a conflicting group that has the maximum degree, where a group of conflicting rules consists of rules with the same IF parts;
3. Linguistic rules from human experts (due to conscious knowledge).

## Step 5. Fuzzy system

In this step of algorithm, the fuzzy system is constructed based on the fuzzy rule base obtained in Step 4 (see [13, 17]).

In the sequel, we present an simple example, with a small amount of inputoutput data, which will illustrate working of the previous procedure and problems that may occur when using this method.

## 3. Example

In order to rate the quality of service offered by the hotel, one hotel booking site measures two components - cleanliness and comfort. Cleanliness takes values from interval $[0,6]$, while comfort takes values from interval $[0,11]$. According to these components, as a result the rating of hotel, which takes values from 1 to 5 , is obtained. The following table presents the rate of the quality that customers specified based on the ratings they gave for cleanliness and comfort:

For this two input - one output space system, we will present how using of automated method for designing fuzzy systems based on learning from example works. Moreover, we will compare the results for certain value, which is obtained when for the same system, we change only the number of membership functions used for presenting the input data.

Table 3.1: Input-output data of example

| Id | Clean | Comfort | Rate |
| :---: | :---: | :---: | :---: |
| 1 | 2.8 | 2 | 2 |
| 2 | 3.9 | 8.2 | 4 |
| 3 | 1.2 | 5 | 2 |
| 4 | 2 | 8.4 | 3 |
| 5 | 5 | 10.3 | 5 |
| 6 | 5 | 9.2 | 4 |
| 7 | 4 | 4 | 3 |
| 8 | 3.7 | 1 | 1 |
| 9 | 4 | 9.8 | 5 |
| 10 | 4 | 8.7 | 4 |
| 11 | 2 | 9 | 3 |
| 12 | 1.3 | 5.7 | 2 |
| 13 | 0.8 | 4.1 | 1 |
| 14 | 3 | 9.4 | 3 |
| 15 | 3.1 | 9.9 | 4 |

For presenting the cleanliness and the comfort the trapezoid functions will be used. The input space is $U_{x}=[0,6] \times[0,11]$. For presenting the rate, singleton functions will be used, and the output space is $U_{y}=\{1, . ., 5\}$.

In the first case, we will have 4 membership functions for the cleanliness and 6 for comfort. Trapezoid functions $A_{1}, A_{2}, A_{3}, A_{4}:[0,6] \rightarrow[0,1]$ for cleanliness:

$$
\begin{aligned}
& A_{1}(x)= \begin{cases}1, & 0 \leqslant x \leqslant 1 ; \\
\frac{2-x}{1}, & 1 \leqslant x \leqslant 2 ; \\
0, & \text { otherwise }\end{cases} \\
& A_{3}(x)= \begin{cases}\frac{x-1}{0.5}, & 1 \leqslant x \leqslant 1.5 \\
1, & 1.5 \leqslant x \leqslant 2.5 \\
\frac{3-x}{0.5}, & 2.5 \leqslant x \leqslant 3 \\
0, & \text { otherwise }\end{cases} \\
& \begin{array}{ll}
\frac{x-2}{1}, & 2 \leqslant x \leqslant 3 ; \\
\frac{5}{2}, & 3 \leqslant x \leqslant 4 ; \\
\frac{5-x}{1}, & 4 \leqslant x \leqslant 5 ; \\
0, & \text { otherwise }
\end{array}
\end{aligned} \quad A_{4}(x)= \begin{cases}\frac{x-3}{2}, & 3 \leqslant x \leqslant 5 ; \\
1, & 5 \leqslant x \leqslant 6 \\
0, & \text { otherwise }\end{cases}
$$

Trapezoid functions $B_{1}, . ., B_{6}:[0,11] \rightarrow[0,1]$ for comfort:

$$
B_{1}\left(x_{2}\right)=\left\{\begin{array}{ll}
1,, & 0 \leqslant x_{2} \leqslant 1 ; \\
\frac{2.5-x}{1.5}, & 1 \leqslant x_{2} \leqslant 2.5 ; \\
0, & \text { otherwise }
\end{array} \quad B_{2}\left(x_{2}\right)= \begin{cases}\frac{x-0.5}{1.5}, & 0.5 \leqslant x_{2} \leqslant 2 \\
1,5 \leqslant x_{2} \leqslant 3 \\
\frac{4.5-x}{1.5}, & 3 \leqslant x_{2} \leqslant 4.5 \\
0, & \text { otherwise }\end{cases}\right.
$$



Fig. 3.1: Membership functions $A_{1}$ (blue), $A_{2}$ (yellow), $A_{3}$ (green) and $A_{4}$ (red).

$$
\begin{aligned}
& B_{3}\left(x_{2}\right)=\left\{\begin{array}{ll}
\frac{x_{2}-2.5}{1.5}, & 2.5 \leqslant x_{2} \leqslant 4 ; \\
1,{ }^{2}-x_{2} & 4 \leqslant x_{2} \leqslant 5 ; \\
\frac{6.5-x_{2}}{1.5}, & 5 \leqslant x_{2} \leqslant 6.5 ; \\
0, & \text { otherwise. }
\end{array} \quad B_{4}\left(x_{2}\right)= \begin{cases}\frac{x_{2}-4.5}{1.5}, & 4.5 \leqslant x_{2} \leqslant 6 ; \\
1, & 6 \leqslant x_{2} \leqslant 7 ; \\
\frac{6.5-x_{2}}{1.5}, & 6.5 \leqslant x \leqslant 8 ; \\
0, & \text { otherwise. }\end{cases} \right. \\
& B_{5}\left(x_{2}\right)=\left\{\begin{array}{ll}
\frac{x_{2}-6.5}{1.5}, & 6.5 \leqslant x_{2} \leqslant 8 ; \\
1, . & 8 \leqslant x_{2} \leqslant 9 ; \\
\frac{10.5-x_{2}}{1.5}, & 9 \leqslant x_{2} \leqslant 10.5 ; \\
0, & \text { otherwise } .
\end{array} \quad B_{6}\left(x_{2}\right)= \begin{cases}\frac{x_{2}-8.5}{1.5}, & 8.5 \leqslant x_{2} \leqslant 10 ; \\
1, & 10 \leqslant x_{2} \leqslant 11 ; \\
0, & \text { otherwise. }\end{cases} \right.
\end{aligned}
$$

FIG. 3.2: Membership functions $B_{1}$ (blue), $B_{2}$ (yellow), $B_{3}$ (green), $B_{4}$ (red), $B_{5}$ (purple) and $B_{6}$ (brown).

In the second case, we will have 2 membership functions for the cleanliness and 3 for comfort. Trapezoid functions $A_{1}^{\prime}, A_{2}^{\prime}:[0,6] \rightarrow[0,1]$ for cleanliness:

$$
A_{1}^{\prime}\left(x_{1}\right)=\left\{\begin{array}{ll}
1, & 0 \leqslant x_{1} \leqslant 2 ; \\
\frac{3.5-x_{1}}{1.5}, & 2 \leqslant x_{1} \leqslant 3.5 ; \\
0, & \text { otherwise } .
\end{array} \quad A_{2}^{\prime}\left(x_{1}\right)= \begin{cases}\frac{x_{1}-1.5}{1.5}, & 1.5 \leqslant x_{1} \leqslant 3 \\
1, & 3 \leqslant x_{1} \leqslant 5 ; \\
0, & \text { otherwise }\end{cases}\right.
$$

Trapezoid functions $B_{1}^{\prime}, . ., B_{3}^{\prime}:[0,11] \rightarrow[0,1]$ for comfort:

$$
B_{1}^{\prime}\left(x_{2}\right)=\left\{\begin{array}{ll}
1, & 0 \leqslant x_{2} \leqslant 3 ; \\
\frac{6-x_{2}}{3}, & 3 \leqslant x_{2} \leqslant 6 ; \\
0, & \text { otherwise }
\end{array} \quad B_{2}^{\prime}\left(x_{2}\right)= \begin{cases}\frac{x_{2}-2}{3}, & 2 \leqslant x_{2} \leqslant 5 \\
1, & 5 \leqslant x_{2} \leqslant 7 \\
\frac{10-x_{2}}{3}, & 7 \leqslant x_{2} \leqslant 10 \\
0, & \text { otherwise }\end{cases}\right.
$$



FIG. 3.3: Membership functions $A_{1}^{\prime}$ (blue) and $A_{2}^{\prime}$ (yellow).

$$
B_{3}^{\prime}\left(x_{2}\right)= \begin{cases}\frac{x_{2}-6}{3}, & 6 \leqslant x_{2} \leqslant 9 \\ 1, & 9 \leqslant x_{2} \leqslant 11 \\ 0, & \text { otherwise }\end{cases}
$$



Fig. 3.4: Membership functions $B_{1}^{\prime}$ (blue), $B_{2}^{\prime}$ (yellow) and $B_{3}^{\prime}$ (green).
For presenting the rate of the hotel, in the both cases, 5 singleton functions $C_{i}: U_{y} \rightarrow\{0,1\}, i \in\{1, . ., 5\}$ will be used:

$$
C_{i}(y)=\left\{\begin{array}{lc}
1, & y=i \\
0, & \text { otherwise }
\end{array}\right.
$$

Fuzzy rule base, constructed from input-output data, in the first case is given by Table 3.2, and in the second case, fuzzy rule base constructed from input-output data is given by Table 3.3.

As we can see from Table 3.3, there is a lot of conflict rules. When we solve all the conflicts, and get rid of double rules, we obtain reduced Model 2., presented in Table 3.4

Fuzzy inference engine used here is Minimum Inference Engine, that is: individualrule based inference with union combination, Mamdani's minimum implication, and $\min$ for all the t-norm operators and max for all the s-norm operators [7, 8]:

$$
O(y)=\max _{l=1}^{M}\left[\sup _{\left(x_{1}, x_{2}\right) \in U_{x}} \min \left(I\left(x_{1}, x_{2}\right), A^{l}\left(x_{1}\right), B^{l}\left(x_{2}\right), C^{l}(y)\right)\right] .
$$

where $M$ is the number of rules. Fuzzifier $I\left(x_{1}, x_{2}\right)$, used here, is the the singleton fuzzifier, i.e. for the given input $\left(x_{1}^{0}, x_{2}^{0}\right)$ :

$$
I\left(x_{1}, x_{2}\right)= \begin{cases}1, & \left(x_{1}, x_{2}\right)=\left(x_{1}^{0}, x_{2}^{0}\right) \\ 0, & \text { otherwise }\end{cases}
$$

Table 3.2: Model 1.

| Rule 1 | IF $x_{1}$ is $A_{2}$ and $x_{2}$ is $B_{2}$ then $y$ is $C_{2}$ |
| :--- | :--- |
| Rule 2 | IF $x_{1}$ is $A_{3}$ and $x_{2}$ is $B_{5}$ then $y$ is $C_{4}$ |
| Rule 3 | IF $x_{1}$ is $A_{1}$ and $x_{2}$ is $B_{3}$ then $y$ is $C_{2}$ |
| Rule 4 | IF $x_{1}$ is $A_{2}$ and $x_{2}$ is $B_{5}$ then $y$ is $C_{3}$ |
| Rule 5 | IF $x_{1}$ is $A_{4}$ and $x_{2}$ is $B_{6}$ then $y$ is $C_{5}$ |
| Rule 6 | IF $x_{1}$ is $A_{4}$ and $x_{2}$ is $B_{5}$ then $y$ is $C_{4}$ |
| Rule 7 | IF $x_{1}$ is $A_{3}$ and $x_{2}$ is $B_{3}$ then $y$ is $C_{3}$ |
| Rule 8 | IF $x_{1}$ is $A_{3}$ and $x_{2}$ is $B_{1}$ then $y$ is $C_{1}$ |
| Rule 9 | IF $x_{1}$ is $A_{3}$ and $x_{2}$ is $B_{6}$ then $y$ is $C_{5}$ |
| Rule 10 | IF $x_{1}$ is $A_{3}$ and $x_{2}$ is $B_{5}$ then $y$ is $C_{4}$ |
| Rule 11 | IF $x_{1}$ is $A_{2}$ and $x_{2}$ is $B_{5}$ then $y$ is $C_{3}$ |
| Rule 12 | IF $x_{1}$ is $A_{1}$ and $x_{2}$ is $B_{4}$ then $y$ is $C_{2}$ |
| Rule 13 | IF $x_{1}$ is $A_{1}$ and $x_{2}$ is $B_{3}$ then $y$ is $C_{1}$ |
| Rule 14 | IF $x_{1}$ is $A_{2}$ and $x_{2}$ is $B_{5}$ then $y$ is $C_{3}$ |
| Rule 15 | IF $x_{1}$ is $A_{2}$ and $x_{2}$ is $B_{6}$ then $y$ is $C_{4}$ |

The outputs obtained by both methods are given in Table 3.5. As we can see from Table 3.5 and Table 3.1, better approximation is obtained by Model 1. For example, in the first row of input output Table 3.1 cleanliness is valued by 2.8 , comfort by 2 and the overall impression rate is 2 , and in the second row of Table 3.2 cleanliness is valued by 2 , comfort by 3 and the overall impression rates are 2 and 1 , by Model 1 and Model 2, respectively. Similarly, fifth row of Table 3.1 corresponds to fourth row of Table 3.4, and again we can see that better approximation is achieved by Model 1. The reason for this lies in the fact that Model 1 consider higher number of fuzzy functions for cleanliness and comfort (in Model 1 there are four fuzzy functions for cleanliness and six fuzzy functions for comfort, while in Model 2 we have only two fuzzy functions for cleanliness and tree for comfort). Therefore Model 1 provides sophisticated and finer fuzzy partition of the universe of the discourse. On the other hand, the Model 2, due to insufficient number of input functions, will never rate the quality of a hotel with a rating of 2 or 5 , which is a serious disadvantage of this model. So, the suggestion is that there must be a lower bound on the number of functions that represent the input data-set.

Table 3.3: Model 2.

| Rule 1 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{2}$ |
| :--- | :--- |
| Rule 2 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{4}$ |
| Rule 3 | IF $x_{1}$ is $A_{1}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{2}$ |
| Rule 4 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{3}$ |
| Rule 5 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{5}$ |
| Rule 6 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{4}$ |
| Rule 7 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{3}$ |
| Rule 8 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{1}^{\prime}$ then $y$ is $C_{1}$ |
| Rule 9 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{5}$ |
| Rule 10 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{4}$ |
| Rule 11 | IF $x_{1}$ is $A_{1}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{3}$ |
| Rule 12 | IF $x_{1}$ is $A_{1}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{1}$ |
| Rule 13 | IF $x_{1}$ is $A_{1}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{1}$ |
| Rule 14 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{3}$ |
| Rule 15 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{4}$ |

Table 3.4: Reduced model 2.

| Rule 1 | IF $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{3}$ |
| :---: | :---: |
| Rule 2 | $x_{1}$ is $A_{1}^{\prime}$ and $x_{2}$ is $B_{2}^{\prime}$ then $y$ is $C_{1}$ |
| Rule 3 | $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{4}$ |
| Rule 4 | $x_{1}$ is $A_{2}^{\prime}$ and $x_{2}$ is $B_{1}^{\prime}$ then $y$ is $C_{1}$ |
| Rule 5 | $x_{1}$ is $A_{1}^{\prime}$ and $x_{2}$ is $B_{3}^{\prime}$ then $y$ is $C_{3}$ |

Table 3.5: Output of the algorithm.

| Id | Input | Model 1 | Model 2 |
| :---: | :---: | :---: | :---: |
| 1 | $(2,3)$ | 2 | 1 |
| 2 | $(2.5,1.9)$ | 2 | 1 |
| 3 | $(3,8)$ | 3 | 4 |
| 4 | $(5,10.6)$ | 5 | 4 |
| 5 | $(4,6)$ | 3 | 3 |

## 4. Conclusion

The paper presents an algorithm for designing fuzzy systems based on learning from examples. The concept uses Mamdani's fuzzy rule systems with a fuzzifier and defuzzifier. Particular attention has been given to the preferences of decision makers involved in the process as experts with extensive practical experience. Based on their opinion, corresponding fuzzy functions that express the importance of attributes in the model are defined. An example to illustrate the process has also been provided. Moreover, through this example, the importance of determining the lower number of functions which represent the input data set is highlighted. In other words, it is shown that the number of these functions significantly influences the quality of the solution.

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Ivana Micić<br>University of Niš<br>Faculty of Sciences and Mathematics<br>Višegradska 33, P. O. Box 224<br>18000 Niš, Serbia<br>ivanajancic84@gmail.com<br>Nada Damljanović<br>University of Kragujevac<br>Faculty of Technical Sciences in Čačak<br>Svetog Save 65, P. O. Box 227<br>32000 Čačak, Serbia<br>nada.damljanovic@ftn.kg.ac.rs<br>Zorana Jančić<br>University of Niš<br>Faculty of Sciences and Mathematics<br>Višegradska 33, P. O. Box 224<br>18000 Niš, Serbia<br>ivanajancic84@gmail.com

# ON DISCRETE WEIGHTED STATISTICAL CONVERGENCE 

Sinan Ercan, Yavuz Altın and Rifat Çolak

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Abstract. In the present paper, the notion of discrete weighted mean method of summability have been extended over the concept of statistical convergence. We have also given the notion of statistical $\left(M, P_{\lambda}\right)$ - summability and $\left[M, P_{\lambda}\right]_{q}$-summability. We have introduced some properties of these modes of convergence.
Keywords: statistical convergence; weighted statistical convergence; statistical ( $\bar{N}, p_{n}$ )summability.

## 1. Introduction

Zygmund [16] introduced the idea of statistical convergence in 1935. Fast and Steinhaus introduced statistical convergence to assign limit to sequences which are not convergent in the usual sense in the same year (see [4],[14]). They used the asymptotic density of a set $A \subset \mathbb{N}$ which is defined as follows:

$$
\delta(A)=\lim _{n \rightarrow \infty} \frac{1}{n}|\{k \leq n: k \in A\}|,
$$

whenever the limit exists. $|\{\}$.$| indicates the cardinality of the enclosed set. A se-$ quence $x=\left(x_{k}\right)$ of numbers is called statistically convergent to a number $\ell$ provided that for $\varepsilon>0$,

$$
\lim _{n \rightarrow \infty} \frac{1}{n}\left|\left\{k \leq n:\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right|=0
$$

In this case, we write $S-\lim _{k \rightarrow \infty} x_{k}=\ell . S$ indicates the set of all statistically convergent sequences. This notion is used an effective tool to resolve many problems in ergodic theory, fuzzy set theory, trigonometric series and Banach spaces. It was studied in summability theory by Kolk et al. (see [8]). Also,. many researchers
studied related topics with summability theory (see[2], [3], [5], [12], [13] ). Furthermore, another type of Cesaro summability was studied by Armitage and Maddox [1].

Moricz and Orhan [11] defined the notion of statistical $\left(\bar{N}, p_{n}\right)$ - summability as: Let $p=\left(p_{k}\right)$ be a sequence of nonnegative real numbers such that $p_{0}>0$, $P_{n}=\sum_{k=0}^{n} p_{k} \rightarrow \infty$ as $n \rightarrow \infty$ and $t_{n}=\frac{1}{P_{n}} \sum_{k=0}^{n} p_{k} x_{k}, n=0,1,2, \ldots$. The sequence $x=\left(x_{k}\right)$ is statistically summable to $\ell$ by the weighted mean method determined by the sequence $\left(p_{k}\right)$ or briefly statistically $\left(\bar{N}, p_{n}\right)$ - summable if

$$
s t-\lim _{n \rightarrow \infty} t_{n}=\ell
$$

$(\bar{N}, s t)$ indicates the set of statistically $\left(\bar{N}, p_{n}\right)$ - summable sequences.

Weighted statistical convergence is introduced by Karakaya and Chisti in [7]. Also Küçükaslan studied this notion in [9]. Then the modified definition is given by Mursaleen et al. in [10] as follows:

A sequence $x=\left(x_{k}\right)$ is weighted statistically convergent (or $S_{\bar{N}}$-convergent) to $\ell$ if for every $\varepsilon>0$, the set $\left\{k \in \mathbb{N}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}$ has weighted density zero, i.e.

$$
\lim _{n \rightarrow \infty} \frac{1}{P_{n}}\left|\left\{k \leq P_{n}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right|=0
$$

where $\mathbb{N}=\{1,2, \ldots\}$. This limit is indicated by $S_{\bar{N}}-\lim _{k \rightarrow \infty} x_{k}=\ell . S(\bar{N})$ denotes the set of these kind of sequences.

The concept of weighted statistical convergence of order $\alpha$ is studied by Ghosal in [6]. Watson introduce the notion of discrete weighted mean method of summability in [15] as follows:

A sequence $\left(x_{k}\right)$ is limitable to $\ell$ by the discrete weighted mean method, if

$$
\tau_{n}=t_{\left[\lambda_{n}\right]}=\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=0}^{\left[\lambda_{n}\right]} p_{k} x_{k} \rightarrow \ell
$$

as $n \rightarrow \infty$ where $\left(\lambda_{n}\right)$ is a real sequence satisfies $1 \leq \lambda_{0}<\lambda_{1}<\ldots \rightarrow \infty$ and $P_{\left[\lambda_{n}\right]}=\sum_{k=0}^{\left[\lambda_{n}\right]} p_{k} \rightarrow \infty$ as $n \rightarrow \infty$ for $p_{0}>0$ and $\left[\lambda_{n}\right]$ denotes the integer part of the number $\lambda_{n}$. The set of these kind sequences denoted by $\left(M_{P_{\lambda}}\right)$.

## 2. Main Results

In this part, first we give the concepts of discrete weighted statistical convergence, $\left[M, P_{\lambda}\right]_{q}$-summability and statistical $\left(M, P_{\lambda}\right)$-summability. Then we establish the relationship between these concepts. The discrete weighted density of a set $K \subseteq \mathbb{N}$ is defined by

$$
\delta_{M}(K)=\lim _{n \rightarrow \infty} \frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: k \in K\right\}\right| .
$$

In particular, if we choose $\lambda_{n}=n$ and $p_{n}=1$, then the discrete weighted density reduces to the natural density.

Throughout this paper $\left(p_{k}\right)$, is a sequence of nonnegative real numbers with $p_{1}>0$ and $P_{\left[\lambda_{n}\right]}=\sum_{k=1}^{\left[\lambda_{n}\right]} p_{k} \rightarrow \infty$ as $n \rightarrow \infty$ and $\lambda=\left(\lambda_{n}\right)$ is a real sequence satisfies $1 \leq \lambda_{1}<\lambda_{2}<\ldots \rightarrow \infty$ as $n \rightarrow \infty$ and we use the notations $\Lambda, \widehat{P}, E\left(\lambda_{n}\right), E(\lambda)$ such as

$$
\begin{gathered}
\Lambda=\left\{\lambda=\left(\lambda_{n}\right): 1 \leq \lambda_{1}<\lambda_{2}<\ldots \rightarrow \infty \text { as } n \rightarrow \infty\right\}, \\
\widehat{P}=\left\{p=\left(p_{k}\right): p_{1}>0, p_{k} \geq 0, k=2,3, \ldots \text { and } P_{\left[\lambda_{n}\right]}=\sum_{k=1}^{\left[\lambda_{n}\right]} p_{k} \rightarrow \infty \text { as } n \rightarrow \infty\right\}, \\
E\left(\lambda_{n}\right)=\left\{k \leq\left[\lambda_{n}\right]: k \in \mathbb{N}\right\}
\end{gathered}
$$

and

$$
E(\lambda)=\left\{\left[\lambda_{n}\right]: n \in \mathbb{N}\right\}=\left\{\left[\lambda_{1}\right],\left[\lambda_{2}\right],\left[\lambda_{3}\right], \ldots\right\}
$$

for a sequence $\lambda=\left(\lambda_{n}\right) \in \Lambda$. Also $A \Delta B=(A \backslash B) \cup(B \backslash A)$.
Definition 2.1 Let $\lambda=\left(\lambda_{n}\right) \in \Lambda$ and $p=\left(p_{k}\right) \in \widehat{P}$ be given. A sequence $x=\left(x_{k}\right)$ is said to be discrete weighted statistically convergent (briefly $S\left(M_{P_{\lambda}}\right)$ convergent) to $\ell$ if the set $\left\{k \in \mathbb{N}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}$ has discrete weighted density zero for every $\varepsilon>0$, i.e.

$$
\lim _{n \rightarrow \infty} \frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right|=0
$$

It is indicated by $S\left(M_{P_{\lambda}}\right)-\lim _{k \rightarrow \infty} x_{k}=\ell . S\left(M_{P_{\lambda}}\right)$ indicates the set of these kind of sequences.

Definition 2.2 Let $\lambda=\left(\lambda_{n}\right) \in \Lambda$ and $p=\left(p_{k}\right) \in \widehat{P}$ be given. A sequence $x=\left(x_{k}\right)$ is called $\left[M, P_{\lambda}\right]_{q}$-summable $(0<q<\infty)$ to the limit $\ell$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q}=0
$$

$\ell$ is said to be $\left[M, P_{\lambda}\right]_{q}$-limit of $x . \quad\left[M, P_{\lambda}\right]_{q}$ indicates the set of these kind of sequences.

Definition 2.3 Let $\lambda=\left(\lambda_{n}\right) \in \Lambda$ and $p=\left(p_{k}\right) \in \widehat{P}$ be given. A sequence $x=\left(x_{k}\right)$ is said to be statistically summable to $\ell$ by the discrete weighted mean method or briefly statistically $\left(M, P_{\lambda}\right)$-summable if

$$
s t-\lim _{n \rightarrow \infty} \tau_{n}=\ell
$$

It is indicated by $\left(M, P_{\lambda}\right)-\lim _{k \rightarrow \infty} x_{k}=\ell$ and we denote by $\left(M, P_{\lambda}\right)$ the set of such sequences.

Note that for any $\left(\lambda_{n}\right)=(n+r) \in \Lambda$, where $0 \leq r<1$ is a fixed number.
(i) $\left[M, P_{\lambda}\right]_{q}$-summable reduces to $\left[\bar{N}, p_{n}\right]_{q}$-summable which is given in [10],
(ii) Discrete weighted statistical convergence is reduced to weighted statistical convergence which is given in [10],
(iii) Statistical $\left(M, P_{\lambda}\right)$-summability is reduced to statistical $\left(\bar{N}, p_{n}\right)$ - summability which is given in [11].

As a result of $(i),(i i)$ and $(i i i)$ we have that $\left[M, P_{\lambda}\right]_{q}$ includes $\left[\bar{N}, p_{n}\right]_{q}, S\left(M_{P_{\lambda}}\right)$ includes $S(\bar{N})$ and $\left(M, P_{\lambda}\right)$ includes $\bar{N}(s t)$, respectively.

We first begin with the following property. In the proof of the following Theorem we use the technique used by Watson in [15].

Theorem 2.4 Let $\lambda=\left(\lambda_{n}\right), \mu=\left(\mu_{n}\right) \in \Lambda$ be given. Then
(i) $\left[M, P_{\lambda}\right]_{q} \subseteq\left[M, P_{\mu}\right]_{q}$ if $E(\mu) \backslash E(\lambda)$ is finite,
(ii) If $p_{k}>0$ for each $k$ and if $\left[M, P_{\lambda}\right]_{q} \subseteq\left[M, P_{\mu}\right]_{q}$ holds, then $E(\mu) \backslash E(\lambda)$ is finite.

Proof. ( $i$ ) Assume that $E(\mu) \backslash E(\lambda)$ is finite. Then we have an integer $n_{0}$ such that $\left\{\left[\mu_{n}\right]: n \geq n_{0}\right\} \subseteq E(\lambda)$. That is, there is an increasing sequence $\left(j_{n}\right)$ of positive integers such that $j_{n} \rightarrow \infty$ and $\left[\mu_{n}\right]=\left[\lambda_{j_{n}}\right]$ for $n \geq n_{0}$. If a sequence $x=\left(x_{n}\right)$ is statistically $\left[M, P_{\lambda}\right]_{q}-$ summable to $\ell$, then we have

$$
\frac{1}{P_{\left[\mu_{n}\right]}} \sum_{k=1}^{\left[\mu_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q}=\frac{1}{P_{\left[\lambda_{j_{n}}\right]}} \sum_{k=1}^{\left[\lambda_{j_{n}}\right]} p_{k}\left|x_{k}-\ell\right|^{q}
$$

for $n \geq n_{0}$, which gives that $x=\left(x_{n}\right)$ is statistically $\left[M, P_{\mu}\right]_{q}$-summable to $\ell$ $(0<q<\infty)$.
(ii) Suppose that $\left[M, P_{\lambda}\right]_{q} \subseteq\left[M, P_{\mu}\right]_{q}$ but $E(\mu) \backslash E(\lambda)$ is infinite. Then there is a strictly increasing sequence $\left(\left[\mu_{n_{j}}\right]\right)$ such that $\left[\mu_{n_{j}}\right] \notin E(\lambda)$, for $j=1,2,3, \ldots$.

Consider that $\tau_{n}=t_{\left[\lambda_{n}\right]}=\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k} x_{k}$. Then define a sequence $\left(\tau_{n}\right)$ as follows.

$$
\tau_{n}=\left\{\begin{array}{ccc}
0 & \text { if } & {\left[\lambda_{n}\right] \neq\left[\mu_{n_{j}}\right]} \\
(-1)^{j} & \text { if } & {\left[\lambda_{n}\right]=\left[\mu_{n_{j}}\right]}
\end{array}\right.
$$

Using the equality $P_{\left[\lambda_{n}\right]} t_{\left[\lambda_{n}\right]}-P_{\left[\lambda_{n}\right]-1} t_{\left[\lambda_{n}\right]-1}=P_{\left[\lambda_{n}\right]} x_{\left[\lambda_{n}\right]}$, we have $x=\left(x_{n}\right) \in$ $\left[M, P_{\lambda}\right]_{q}$ since $t_{\left[\lambda_{n}\right]}=0$ for all $n$. But the sequence is not in $\left[M, P_{\mu}\right]_{q}$.

We have the below results from Theorem 2.4.
Corollary 2.5 Let $\lambda=\left(\lambda_{n}\right), \mu=\left(\mu_{n}\right) \in \Lambda$ be given and assume that $p_{k}>0$ for every $k$. Then
(i) $\left[M, P_{\lambda}\right]_{q} \subseteq\left[M, P_{\mu}\right]_{q}$ if and only if $E(\mu) \backslash E(\lambda)$ is a finite set.
(ii) $\left[M, P_{\lambda}\right]_{q}=\left[M, P_{\mu}\right]_{q}$ if and only if $E(\lambda) \Delta E(\mu)$ is finite set.

Corollary 2.6 Let $\lambda=\left(\lambda_{n}\right), \mu=\left(\mu_{n}\right) \in \Lambda$ be given and assume that $p_{k}>0$ for every $k$. Then
(i) $\left(M, P_{\lambda}\right) \subseteq\left(M, P_{\mu}\right)$ if and only if $E(\mu) \backslash E(\lambda)$ is a finite set.
(ii) $\left(M, P_{\lambda}\right)=\left(M, P_{\mu}\right)$ if and only if $E(\lambda) \Delta E(\mu)$ is a finite set.

The following property has been given formerl, but it is also seen clearly from Theorem 2.4 by taking $\lambda=\left(\lambda_{n}\right)=(n)$.

Corollary 2.7 For any $\left(\mu_{n}\right) \in \Lambda$ the inclusion $\left[\bar{N}, p_{n}\right]_{q} \subseteq\left[M, P_{\mu}\right]_{q}$ is satisfied, where $0<q<\infty$.

Corollary 2.8 For any $\left(\lambda_{n}\right) \in \Lambda\left[M, P_{\lambda}\right]_{q} \subseteq\left[\bar{N}, p_{n}\right]_{q}$ is satisfied if $\{1,2,3, \ldots\} \backslash E(\lambda)$ is finite.

The following results can be obtained from Corollary 2.5 and Corollary 2.6.
Corollary 2.9 (i) $\left(\bar{N}, p_{n}\right) \subseteq\left(M, P_{\mu}\right)$ for any $\mu=\left(\mu_{n}\right) \in \Lambda$.
(ii) $\left(M, P_{\lambda}\right) \subseteq\left(\bar{N}, p_{n}\right)$ if $\{1,2,3, \ldots\} \backslash E(\lambda)$ is a finite set.

Theorem 2.10 If a sequence $x=\left(x_{k}\right)$ is $\left(M_{P_{\lambda}}\right)$-summable to $\ell$, then it is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$. The inverse implication need not be true.

Proof. Let $x \in\left(M_{P_{\lambda}}\right)$ and define the set $K_{P_{\lambda}}(\varepsilon)=\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}$ for $\varepsilon>0$. Hence the inequality which we have

$$
\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|=\frac{1}{P_{\left[\lambda_{n}\right]}}\left(\sum_{k \in K_{P_{\lambda}}(\varepsilon)}+\sum_{k \notin K_{P_{\lambda}}(\varepsilon)}\right) p_{k}\left|x_{k}-\ell\right|
$$

$$
\begin{aligned}
& \geq \frac{1}{P_{[\lambda n]}} \sum_{k \in K_{P_{\lambda}}(\varepsilon)} p_{k}\left|x_{k}-\ell\right| \\
& \geq \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k \in K_{P_{\lambda}}(\varepsilon)} \varepsilon \\
& =\varepsilon \cdot \frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right|
\end{aligned}
$$

This implies that $x \in S\left(M_{P_{\lambda}}\right)$. To see the converse implication is not true, consider $\lambda_{n}=n, p_{k}=1$ for all $k$ and define $x=\left(x_{k}\right)$ by

$$
x_{k}=\left\{\begin{array}{cl}
m^{3}, & k=m^{2} \\
0, & k \neq m^{2}
\end{array} \quad m=1,2, \ldots\right.
$$

Now we have

$$
\lim _{n \rightarrow \infty} \frac{1}{n}\left|\left\{k \leq n: 1 .\left|x_{k}-0\right| \geq \varepsilon\right\}\right|=0 \text { and } \lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n} 1\left|x_{k}-0\right|=\infty
$$

This means $x \in S\left(M_{P_{\lambda}}\right)$ but $x \notin\left(M_{P_{\lambda}}\right)$.
Theorem 2.11 Assume that $p_{n} \geq 1$ for all $n$ and

$$
\begin{equation*}
1 \leq \lim _{n \rightarrow \infty} \frac{P_{\left[\lambda_{n}\right]}}{n}<\infty \tag{2.1}
\end{equation*}
$$

holds. Then $x \in S$ if $x \in S\left(M_{P_{\lambda}}\right)$. The opposite case is not true.

Proof. Suppose that $p_{n} \geq 1$ and (2.1) holds for all $n \in \mathbb{N}$. Let $x \in S\left(M_{P_{\lambda}}\right)$, then we have

$$
\begin{aligned}
\frac{1}{n}\left|\left\{k \leq n:\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right| & \leq \frac{1}{n}\left|\left\{k \leq n: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right| \\
& \leq \frac{P_{\left[\lambda_{n}\right]}}{n} \frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right|
\end{aligned}
$$

for $\varepsilon>0$. As $n \rightarrow \infty$ we obtain $x \in S$.

To see the opposite case is not true, consider the sequence $x=\left(x_{k}\right)$ defined by

$$
x_{k}=\left\{\begin{array}{cl}
1 & \text { if } k=m^{2} \\
\frac{1}{\sqrt{k}} & \text { if } k \neq m^{2}
\end{array} \quad m \in \mathbb{N}\right.
$$

The sequence $x=\left(x_{k}\right)$ is a statistically convergent to 0 , but it is not discrete weighted statistical convergent to 0 , while $p_{k}=k$ for $k \in \mathbb{N}$.

The sufficient condition to be true the converse implication is given the following theorem.

Theorem 2.12 Assume that $p_{n}<1$ and

$$
\begin{equation*}
1 \leq \lim _{n \rightarrow \infty} \frac{n}{P_{\left[\lambda_{n}\right]}}<\infty \tag{2.2}
\end{equation*}
$$

for all $n \in \mathbb{N}$. If $x \in S$, then $x \in S\left(M_{P_{\lambda}}\right)$.

Proof. Suppose that $p_{n}<1$ and (2.2) holds for all $n \in \mathbb{N}$. Let $x \in S$ then we have

$$
\begin{aligned}
\frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right| & \leq \frac{1}{P_{\text {(גn] }}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}:\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right| \\
& \leq \frac{n^{n}}{P_{\left[\lambda_{n}\right]}} \frac{1}{n}\left|\left\{k \leq n:\left|x_{k}-\ell\right| \geq \varepsilon\right\}\right|
\end{aligned}
$$

for any $\varepsilon>0$. Hence we obtain the desired result as $n \rightarrow \infty$.

Theorem 2.13 $S\left(M_{P_{\lambda}}\right)$-lim of an $S\left(M_{P_{\lambda}}\right)$-convergent sequence is unique.

Proof. Suppose the sequence $x=\left(x_{k}\right)$ is $S\left(M_{P_{\lambda}}\right)$-convergent both to $\ell_{1}$ and $\ell_{2}$. If possible let $\ell_{1} \neq \ell_{2}$ and choose $\varepsilon=\frac{1}{2}\left|\ell_{1}-\ell_{2}\right|>0$ and $p_{k}>c>0$ for all $k$. Then

$$
\begin{aligned}
1 & \leq \frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|\ell_{1}-\ell_{2}\right| \geq \varepsilon c\right\}\right| \\
& \leq \frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell_{1}\right| \geq \frac{\varepsilon c}{2}\right\}\right|+\frac{1}{P_{\left[\lambda_{n}\right]}}\left|\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell_{2}\right| \geq \frac{\varepsilon c}{2}\right\}\right| .
\end{aligned}
$$

This is impossible because right hand side tends to 0 as $n \rightarrow \infty$. Hence we have desired result $\ell_{1}=\ell_{2}$.

Theorem 2.14 Let the sequence $\left(p_{k}\left|\left(x_{k}-\ell\right)\right|\right)$ be bounded. Then $x=\left(x_{k}\right)$ is statistically $\left(M_{P_{\lambda}}\right)$ - summable to $\ell$ if it is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$, but the opposite case is not true.

Proof. Suppose $p_{k}\left|\left(x_{k}-\ell\right)\right| \leq T$ for every $k$, for some constant $T$ and assume that the sequence $x=\left(x_{k}\right)$ is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$. We have

$$
\begin{aligned}
\left|\tau_{n}-\ell\right| & =\left\lvert\, \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k} x_{k}-\ell\right.
\end{aligned}\left|, \quad \begin{array}{|l}
\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left(x_{k}-\ell\right)
\end{array}\right|
$$

Theorem 2.15 Let the sequence $\left(p_{k}\left|\left(x_{k}-\ell\right)\right|\right)$ be bounded. Then $x=\left(x_{k}\right)$ is statistically $\left(M_{P_{\lambda}}\right)$ - summable to $\ell$ if it is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$, but the opposite case is not true.

Proof. Suppose $p_{k}\left|\left(x_{k}-\ell\right)\right| \leq T$ for every $k$, for some constant $T$ and assume that the sequence $x=\left(x_{k}\right)$ is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$. We have

$$
\begin{aligned}
& \left.\begin{aligned}
\left|\tau_{n}-\ell\right| & =\left|\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k} x_{k}-\ell\right| \\
& =\left\lvert\, \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left(x_{k}-\ell\right)\right.
\end{aligned} \right\rvert\, \\
& =\left|\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n]}\right]} p_{k}\left(x_{k}-\ell\right)+\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}^{c}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left(x_{k}-\ell\right)\right| \\
& \leq\left|\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left(x_{k}-\ell\right)\right|+\left|\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}^{c}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left(x_{k}-\ell\right)\right| \\
& \leq \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|\left(x_{k}-\ell\right)\right|+\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}^{c}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|\left(x_{k}-\ell\right)\right| \\
& \leq \frac{1}{P_{[\lambda n]}} \cdot T \cdot\left|K_{P_{\lambda}}(\varepsilon)\right|+\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k \in K_{P_{\lambda}}^{c}} \varepsilon \\
& =\frac{1}{P_{\left[\lambda_{n}\right]}} \cdot T \cdot\left|K_{P_{\lambda}}(\varepsilon)\right|+\varepsilon \frac{\left|K_{P_{\lambda}}^{c}(\varepsilon)\right|}{P_{\left[\lambda_{n}\right]}} \rightarrow 0
\end{aligned}
$$

as $n \rightarrow \infty$, where $K_{P_{\lambda}}(\varepsilon)=\left\{k \leq P_{\left[\lambda_{n}\right]}: p_{k}\left|x_{k}-\ell\right| \geq \varepsilon\right\}$. This means that $\tau_{n} \rightarrow$ $\ell$ as $n \rightarrow \infty$. That is, $x$ is $\left(M_{P_{\lambda}}\right)$-summable to $\ell$ and hence it is statistically ( $M_{P_{\lambda}}$ )-summable to $\ell$.

To see that the opposite case is not true, let $p_{k}=1$ for every $k \in \mathbb{N}$. Consider the sequence $x=\left(x_{k}\right)$ defined by

$$
x_{k}=\left\{\begin{array}{cc}
1 & \text { if } k=m^{2}-m, m^{2}-m+1, \ldots, m^{2}-1 \\
-m & \text { if } k=m^{2} \\
0 & \text { otherwise }
\end{array}\right.
$$

where $m=2,3,4, \ldots$. Then we have, for $s=0,1,2, \ldots, m-1 ; m=2,3, \ldots$

$$
\tau_{n}=t_{\left[\lambda_{n}\right]}=\frac{1}{\left[\lambda_{n}\right]+1} \sum_{k=0}^{\left[\lambda_{n}\right]} x_{k}=\left\{\begin{array}{cc}
\frac{s+1}{\left[\lambda_{n}\right]+1} & \text { if }\left[\lambda_{n}\right]=m^{2}-m+s \\
0 & \text { otherwise }
\end{array}\right.
$$

It is easy to see that $\lim _{n \rightarrow \infty} \tau_{n}=0$ and hence $s t-\lim _{n \rightarrow \infty} \tau_{n}=0$, i.e. $x=\left(x_{k}\right)$ is statistically $\left(M_{P_{\lambda}}\right)$-summable to 0 . On the other hand $s t-\lim _{k \rightarrow \infty} \inf x_{k}=0$ and $s t-\lim _{k \rightarrow \infty} \sup x_{k}=1$, because

$$
\delta\left(\left\{k: k=m^{2}, m=1,2,3, \ldots\right\}\right)=0
$$

$$
\delta\left(\left\{k: k \neq m^{2}-m, m^{2}-m+1, \ldots, m^{2}-1, m ; m=2,3, \ldots\right\}\right) \neq 0
$$

and

$$
\delta\left(\left\{k: k=m^{2}-m, m^{2}-m+1, \ldots, m^{2}-1 ; m=2,3, \ldots\right\}\right) \neq 0 .
$$

Hence $x=\left(x_{k}\right)$ is not $S\left(M_{P_{\lambda}}\right)$-convergent.
Theorem 2.16 Let a sequence $x=\left(x_{k}\right)$ be $\left[M_{P_{\lambda}}\right]_{q}$-summable to $\ell$. If $0<q<1$ and $0 \leq\left|x_{k}-\ell\right|<1$ or $1 \leq q<\infty$ and $1 \leq\left|x_{k}-\ell\right|<\infty$, then $x=\left(x_{k}\right)$ is $S\left(M_{P_{\lambda}}\right)$-statistically convergent to $\ell$.

Proof. We have $p_{k}\left|x_{k}-\ell\right|^{q} \geq p_{k}\left|x_{k}-\ell\right|$ for both cases. Then

$$
\begin{aligned}
\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} & \geq \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right| \\
& \geq \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right| \\
& \geq \frac{1}{P_{[\lambda n]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} \\
& =\varepsilon \frac{\left|K_{P_{\lambda}}(\varepsilon)\right|}{P_{\left[\lambda_{n}\right]}} .
\end{aligned}
$$

Since $\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} \rightarrow 0$ as $n \rightarrow \infty$ we have $\lim _{n \rightarrow \infty} \frac{1}{P_{\left[\lambda_{n}\right]}}\left|K_{P_{\lambda}}(\varepsilon)\right|=0$. This means that $x=\left(x_{k}\right)$ is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$.

Theorem 2.17 Let the sequence $\left(p_{k}\left|\left(x_{k}-\ell\right)\right|\right)$ be bounded and let a sequence $x=\left(x_{k}\right)$ be $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$. If $0<q<1$ and $0 \leq T<\infty$ or $1 \leq q<\infty$ and $0 \leq T<1$, then $x=\left(x_{k}\right)$ is $\left[M_{P_{\lambda}}\right]_{q}$-summable to $\ell$.

Proof. Assume that $x=\left(x_{k}\right)$ is $S\left(M_{P_{\lambda}}\right)$-convergent to $\ell$. Since $p_{k}\left|x_{k}-\ell\right| \leq T$ $(k=1,2, \ldots)$ for some $T \geq 0$, we have

$$
\begin{aligned}
\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{k=1}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} & =\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \notin K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q}+\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} \\
& =s_{1}\left(\left[\lambda_{n}\right]\right)+s_{2}\left(\left[\lambda_{n}\right]\right)
\end{aligned}
$$

where

$$
s_{1}\left(\left[\lambda_{n}\right]\right)=\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\ k \notin K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q}
$$

and

$$
s_{2}\left(\left[\lambda_{n}\right]\right)=\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\ k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} .
$$

Now we have

$$
\begin{aligned}
s_{1}\left(\left[\lambda_{n}\right]\right) & =\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \notin K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} \\
& \leq \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \notin K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|=\varepsilon \frac{1}{P_{\left[\lambda_{n}\right]}}\left|K_{P_{\lambda}}^{c}(\varepsilon)\right| \rightarrow 0
\end{aligned}
$$

and

$$
\begin{aligned}
s_{2}\left(\left[\lambda_{n}\right]\right) & =\frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right|^{q} \leq \frac{1}{P_{\left[\lambda_{n}\right]}} \sum_{\substack{k=1 \\
k \in K_{P_{\lambda}}(\varepsilon)}}^{\left[\lambda_{n}\right]} p_{k}\left|x_{k}-\ell\right| \\
& \leq\left(\sup _{k} p_{k}\left|x_{k}-\ell\right|\right)\left(\left|K_{P_{\lambda}}(\varepsilon)\right| / P_{\left[\lambda_{n}\right]}\right) \leq T\left|K_{P_{\lambda}}(\varepsilon)\right| / P_{\left[\lambda_{n}\right]} \rightarrow 0
\end{aligned}
$$

as $n \rightarrow \infty$. Hence $\left(x_{k}\right)$ is $\left[M_{P_{\lambda}}\right]_{q}$-summable to $\ell$.

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Sinan Ercan<br>Department of Mathematics, Firat University Elazig, Türkiye sinanercan45@gmail.com

Yavuz Altın<br>Department of Mathematics, Firat University, Elazig, Türkiye yaltin23@yahoo.com

Rifat Çolak
Department of Mathematics
Firat University, Elazig, Türkiye
rftcolak@gmail.com

# APPLICATIONS OF MATRIX TRANSFORMATIONS TO ABSOLUTE SUMMABILITY 

Mehmet Ali Sarıg̈l

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Abstract. Rhoades and Savaş [6], [11] established necessary conditions for inclusions of the absolute matrix summabilities under additional conditions. In this paper, we determine necessary or sufficient conditions for some classes of infinite matrices, and using this, we get necessary or sufficient conditions for more general absolute summabilities applied to all matrices.
Keywords: matrix summability; infinite matrices; Cesàro matrices; triangular matrix.

## 1. Introduction

Let $X$ and $Y$ be two sequence spaces of the space $\omega$, the set of all sequences with real or complex terms. Let $A=\left(a_{n v}\right)$ be an infinite matrix of complex numbers. By $A(x)=\left(A_{n}(x)\right)$, we denote the $A$-transform of the sequence $x=\left(x_{v}\right)$, i.e.,

$$
A_{n}(x)=\sum_{v=0}^{\infty} a_{n v} x_{v}
$$

provided that the series are convergent for $v, n \geq 0$. If $A(x) \in Y$ for all $x \in X$, then $A$ is called a matrix transformation from $X$ into $Y$, and denoted by $(X, Y)$.

In many cases, since an infinite matrix can be considered as a linear operator between two sequence spaces, the theory of matrix transformations in sequence spaces has aroused interest for many years, of which purpose is to provide the necessary and sufficient conditions for a matrix to map a sequence space into another.
$X$ is called a $B K$-space, if it is a Banach space on which all coordinate functionals defined by $p_{n}(x)=x_{n}$ are continuous.

Let $\Sigma a_{v}$ be a given infinite series with $n$-th partial sum $s_{n}$ and let $\left(\gamma_{n}\right)$ be a sequence of nonnegative numbers. By $\left(A_{n}(s)\right)$, we denote the $A$-transform of the sequence $s=\left(s_{n}\right)$. The series $\Sigma x_{v}$ is said to be summable $\left|A, \gamma_{n}\right|_{k}, k \geq 1$, if (see [7])

$$
\begin{equation*}
\sum_{n=1}^{\infty} \gamma_{n}^{k-1}\left|A_{n}(s)-A_{n-1}(s)\right|^{k}<\infty \tag{1.1}
\end{equation*}
$$

Note that, for $\gamma_{n}=n,\left|A, \gamma_{n}\right|_{k}=|A|_{k}$ [12], Also, if $A$ is chosen as the matrices of the weighted mean $\left(R, p_{n}\right)$ (resp. $\left.\gamma_{n}=P_{n} / p_{n}\right)$ and Cesàro mean $(C, \alpha)$ together with $\gamma_{n}=n$, then, it reduces to the summabilities $\left|R, p_{n}\right|_{k}[8]\left(\right.$ resp. $\left.\left|\bar{N}, p_{n}\right|_{k} \quad[1]\right)$ and $|C, \alpha|_{k}$ [2], respectively. By the weighted and Cesàro matrices we mention

$$
a_{n v}=\left\{\begin{array}{cc}
\frac{p_{v}}{P_{n}}, & 0 \leq v \leq n \\
0, & v>n
\end{array}\right.
$$

and

$$
a_{n v}=\left\{\begin{array}{cc}
\frac{A_{n-v}^{\alpha-1}}{A_{n}^{\alpha}}, \quad 0 \leq v \leq n \\
0, & v>n
\end{array}\right.
$$

respectively, where $\left(p_{n}\right)$ is a sequence of positive numbers with $P_{n}=p_{0}+p_{1}+\ldots+$ $p_{n} \rightarrow \infty$, and

$$
\begin{gathered}
A_{n}^{\alpha}=\frac{(\alpha+1)(\alpha+2) \cdots(\alpha+n)}{n!}, n \geq 1, A_{0}^{\alpha}=1 \\
\left|A_{n}^{\alpha}\right| \leq A(\alpha) n^{\alpha} \text { for all } \alpha \\
A_{n}^{\alpha} \geq A(\alpha) n^{\alpha} \text { and } A_{n}^{\alpha}>0 \text { for } \alpha>-1 .
\end{gathered}
$$

Let $A=\left(a_{n v}\right)$ be a lower triangular matrix, we derive the matrices $\bar{A}=\left(\bar{a}_{n v}\right)$ and $\widehat{A}=\left(\widehat{a}_{n v}\right)$ from the matrix $A$ as follows:

$$
\begin{aligned}
\bar{a}_{00} & =\widehat{a}_{00}=a_{00} \\
\bar{a}_{n v} & =\sum_{r=v}^{n} a_{n r} ; n, v=0,1, \ldots \\
\widehat{a}_{n v} & =\bar{a}_{n v}-\bar{a}_{n-1, v}, \quad \bar{a}_{n-1, n}=0 .
\end{aligned}
$$

Then, $\widehat{A}$ is a triangular matrix and has unique inverse which is also triangular (see [13]). We will denote its inverse $\widehat{A}^{\prime}$. Hence, it can be written that

$$
A_{n}(x)=\sum_{v=0}^{n} a_{n v} s_{v}=\sum_{r=0}^{n}\left(\sum_{v=r}^{n} a_{n v}\right) x_{r}=\sum_{v=0}^{n} \bar{a}_{n v} x_{v}
$$

and

$$
\begin{equation*}
\widehat{A}_{n}(x)=A_{n}(x)-A_{n-1}(x)=\sum_{v=0}^{n}\left(\bar{a}_{n v}-\bar{a}_{n-1, v}\right) x_{v}=\sum_{v=0}^{n} \widehat{a}_{n v} x_{v} \tag{1.2}
\end{equation*}
$$

which means that the summability $\left|A, \gamma_{n}\right|_{k}$ is equivalent to

$$
\begin{equation*}
\sum_{n=0}^{\infty} \gamma_{n}^{k-1}\left|\widehat{A}_{n}(x)\right|^{k}<\infty \tag{1.3}
\end{equation*}
$$

By $|\gamma A|_{k}$, we define the set of all series summable by $\left|A, \gamma_{n}\right|_{k}$. Then, a series $\Sigma x_{v}$ is summable $\left|A, \gamma_{n}\right|_{k}$ iff $x=\left(x_{v}\right) \in|\gamma A|_{k}$, i.e.,

$$
\begin{equation*}
|\gamma A|_{k}=\left\{x=\left(x_{v}\right): \widetilde{A}(x)=\left(\widetilde{A}_{n}(x)\right) \in \ell_{k}\right\} \tag{1.4}
\end{equation*}
$$

where $\widetilde{A}_{n}(x)=\gamma_{n}^{1-1 / k} \widehat{A}_{n}(x)$ for all $n \geq 0$ and $\ell_{k}$ is the set of all $k$-absolutely convergent series.

We note that, since $\widetilde{A}=\left(\widetilde{a}_{n v}\right)$ is a triangle matrix, it is routine to show that $|\gamma A|_{k}$ is a $B K$ - space if normed by

$$
\begin{equation*}
\|x\|_{|\gamma A|_{k}}=\|\widetilde{A}(x)\|_{\ell_{k}}, \quad 1 \leq k<\infty \tag{1.5}
\end{equation*}
$$

Dealing with the absolute weighted mean summability of infinite series, Bor and Thorpe [1] established sufficient conditions in order that all $\left|\bar{N}, p_{n}\right|_{k}$ summable series is also summable $\left|\bar{N}, q_{n}\right|_{k}$, and conversely. The author [10] showed that Bor and Thorpe's conditions are not only sufficient but also necessary for the conclusion. Also, these results of the author [10] were extended by Rhoades and Savaş [6] using a triangle matrix instead of weighted mean matrix as follows.

Theorem 1.1. Let $1<k \leq s<\infty,\left(p_{n}\right)$ be a sequence satisfying

$$
\begin{equation*}
\sum_{n=v+1}^{\infty} n^{k-1}\left(\frac{p_{n}}{P_{n} P_{n-1}}\right)^{k}=O\left(\frac{1}{P_{n}^{k}}\right) \tag{1.6}
\end{equation*}
$$

Let $B$ be a lower triangular matrix. Then, necessary conditions for $\Sigma x_{v}$ summable $\left|\bar{N}, p_{n}\right|_{k}$ to imply $\Sigma x_{v}$ is summable $|B|_{s}$ are

$$
\begin{gathered}
\frac{P_{v}\left|b_{v v}\right|}{p_{v}}=O\left(v^{1 / s-1 / k}\right) \\
\sum_{n=v+1}^{\infty} n^{s-1}\left|\Delta_{v} \widehat{b}_{n v}\right|^{s}=O\left(v^{s-s / k} \frac{p_{v}}{P_{v}}\right) \\
\sum_{n=v+1}^{\infty} n^{s-1}\left|\widehat{b}_{n, v+1}\right|^{s}=O(1)
\end{gathered}
$$

This result has also been extended by Savaş [11] to the matrix methods as follows

Theorem 1.2. Let $1<k \leq s<\infty, A$ and $B$ be two lower triangular matrices. $A$ satisfying

$$
\begin{equation*}
\sum_{n=v+1}^{\infty} n^{k-1}\left|\Delta_{v} \widehat{a}_{n v}\right|^{k}=O\left(\left|a_{v v}\right|^{k}\right) \tag{1.7}
\end{equation*}
$$

Then necessary conditions for $\Sigma x_{v}$ summable $|A|_{k}$ to imply $\Sigma x_{v}$ is summable $|B|_{s}$ are

$$
\begin{gathered}
\left|b_{v v}\right|=O\left(v^{1 / s-1 / k}\left|a_{v v}\right|\right) \\
\sum_{n=v+1}^{\infty} n^{s-1}\left|\Delta_{v} \widehat{b}_{n v}\right|^{s}=O\left(v^{s-s / k}\left|a_{v v}\right|^{s}\right)
\end{gathered}
$$

and

$$
\sum_{n=v+1}^{\infty} n^{s-1}\left|\widehat{b}_{n, v+1}\right|^{s}=O\left(\sum_{n=v+1}^{\infty} n^{k-1}\left|\widehat{a}_{n, v+1}\right|^{k}\right)^{s / k} .
$$

## 2. Main results

We note that Theorem 1.1 and Theorem 1.2 give necessary conditions for the triangle matrices under the conditions (1.6) and (1.7). In the present paper, we determine necessary or sufficient conditions for a matrix $T \in\left(|\gamma A|_{k},|\phi B|_{s}\right), 1 \leq k \leq s<\infty$. Also, in the special case, we get some more general results that do not include the conditions (1.6) and (1.7). More precisely, we give the following theorems.

Theorem 2.1. Let $A, B$ be infinite triangle matrix and $T$ be any infinite matrix of complex numbers. Further, let $\left(\gamma_{n}\right)$ and $\left(\phi_{n}\right)$ be two sequences of positive numbers. Then, the necessary conditions for $T \in\left(|\gamma A|_{k},|\phi B|_{s}\right), 1<k \leq s<\infty$, are

$$
\begin{equation*}
\bar{l}_{n r}=\gamma_{r}^{-1 / k^{*}} \sum_{i=r}^{\infty} t_{n i} \widehat{a}_{i r}^{\prime} \text { converges for } n, r \geq 0 \tag{2.1}
\end{equation*}
$$

$$
\begin{align*}
& \sup _{m} \sum_{v=0}^{m} \frac{1}{\gamma_{r}}\left|\sum_{v=r}^{m} t_{n v} \widehat{a}_{v r}^{\prime}\right|^{k^{*}}<\infty \text { for } n, r \geq 0  \tag{2.2}\\
& \sum_{n=m}^{\infty} \phi_{n}^{s-1}\left|\sum_{v=0}^{n} \sum_{i=m}^{\infty} \widehat{b}_{n v} t_{v i} \widehat{a}_{i m}^{\prime}\right|^{s}=O\left(\gamma_{m}^{s / k^{*}}\right),
\end{align*}
$$

where $k^{*}$ is the conjugate of $k$, i.e., $k^{*}=k /(k-1)$.

Theorem 2.2. Let $A, B$ be infinite triangle matrix and $T$ be any infinite matrix of complex numbers. Further, let $\left(\phi_{n}\right)$ be a sequences of positive numbers. Then, the necessary and sufficient conditions for $T \in\left(|A|,|\phi B|_{s}\right), 1=k \leq s<\infty$, are

$$
\begin{equation*}
\bar{l}_{n r}=\sum_{i=r}^{\infty} t_{n i} \widehat{a}_{i r}^{\prime} \text { converges for all } n, r \geq 0 \tag{2.4}
\end{equation*}
$$

$$
\begin{align*}
& \sup _{m, r}\left|\sum_{v=r}^{m} t_{n v} \widehat{a}_{v r}^{\prime}\right|<\infty  \tag{2.5}\\
& \sum_{n=0}^{\infty}\left|\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{l}_{v r}\right|^{s}=O(1) . \tag{2.6}
\end{align*}
$$

Note that for $1<k \leq s<\infty$, the characterization of the class of all matrices $\left(\ell_{k}, \ell_{s}\right)$ are not known. Hence one can not expect to get a set of necessary and sufficient conditions for Theorem 2.1.

We require the following lemmas for the proof of our theorems.
Lemma A. Let $X$ and $Y$ be $B K$ spaces, and $A$ be an infinite matrix of complex numbers. If $A$ is a matrix transformation from $X$ into $Y$, i.e., $A \in(X, Y)$, then it is a bounded linear operator [13].

Lemma B. Let $1<k<\infty$ and $A$ be an infinite matrix of complex numbers. Then
a-) $A \in(\ell, c)$ iff
$i-) \lim _{n} a_{n v}$ exists for all $v \geq 0$, and $\left.i i-\right) \sup _{n, v}\left|a_{n v}\right|<\infty$,
b-) $A \in\left(\ell_{k}, c\right)$ iff

$$
i-)(i) \text { is satisfied, and } \quad i i-) \sup _{n} \sum_{v=0}^{\infty}\left|a_{n v}\right|^{k^{*}}<\infty
$$

where $c$ is the set of all convergent sequences, and $1 / k+1 / k^{*}=1[13]$.
Lemma C. Let $1 \leq s<\infty$ and $A$ be an infinite matrix. Then $A \in\left(\ell_{1}, \ell_{s}\right)$ iff

$$
\sup _{v} \sum_{n=0}^{\infty}\left|a_{n v}\right|^{s}<\infty
$$

where $\ell_{s}$ is the set of all $s$ - absolutely convergent sequences [3].
Proof of the Theorem 2.1. Let $1<k \leq s<\infty$. Suppose, $T \in\left(|\gamma A|_{k},|\phi B|_{s}\right)$. Then, $T(x)$ exists and $T(x) \in|\phi B|_{s}$ for all $x \in|\gamma A|_{k}$. Now, $x \in|\gamma A|_{k}$ iff $y=\widetilde{A}(x) \in$
$\ell_{k}$, where $y_{n}=\widetilde{A}_{n}(x)=\gamma_{n}^{1 / k^{*}} \widehat{A}_{n}(x)$, and $\widehat{A}_{n}(x)$ is defined by (1.2). By the inverse of (1.2), we have

$$
x_{n}=\sum_{r=0}^{n} \widehat{a}_{n r}^{\prime} \widehat{A}_{r}(x)=\sum_{r=0}^{n} \widehat{a}_{n r}^{\prime} \gamma_{r}^{-1 / k^{*}} y_{r}
$$

and so

$$
\begin{aligned}
\sum_{v=0}^{m} t_{n v} x_{v} & =\sum_{v=0}^{m} t_{n v} \sum_{r=0}^{v} \widehat{a}_{v r}^{\prime} \gamma_{r}^{-1 / k^{*}} y_{r} \\
& =\sum_{r=0}^{m}\left(\gamma_{r}^{-1 / k^{*}} \sum_{v=r}^{m} t_{n v} \widehat{a}_{v r}^{\prime}\right) y_{r}=\sum_{r=0}^{\infty} l_{m r}^{(n)} y_{r} \\
& =L_{m}^{(n)}(y)
\end{aligned}
$$

where

$$
l_{m r}^{(n)}=\left\{\begin{array}{cc}
\gamma_{r}^{-1 / k^{*}} \sum_{v=r}^{m} t_{n v} \widehat{a}_{v r}^{\prime}, & 0 \leq r \leq m \\
0, & r>m
\end{array}\right.
$$

This implies that $T(x)$ exists for all $x \in|\gamma A|_{k}$ iff $L^{(n)}(y)$ exists for $y \in \ell_{k}$, or equivalently, $L^{(n)}=\left(l_{m r}^{(n)}\right) \in\left(\ell_{k}, c\right)$. So, it follows from Lemma $B$ that $T(x)$ exists iff (2.1) and (2.2) are satisfied. Further,

$$
\begin{aligned}
T_{n}(x) & =\sum_{v=0}^{\infty} t_{n v} x_{v}=\sum_{r=0}^{\infty} \lim _{m \rightarrow \infty} l_{m r}^{(n)} y_{r} \\
& =\sum_{r=0}^{\infty} \bar{l}_{n r} y_{r}=\bar{L}_{n}(y)
\end{aligned}
$$

which means $T(x)=\bar{L}(y)$. On the other hand, since $x \in|\phi B|_{s}$ iff $\widetilde{B}_{n}(x) \in \ell_{s}$, $T(x) \in|\phi B|_{s}$ iff $\widetilde{B}_{n}(T(x)) \in \ell_{s}$, i.e., $C(y) \in \ell_{s}$, where

$$
c_{n r}=\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{l}_{v r} \text { for } n, r \geq 0
$$

because, for each $n \geq 0$,

$$
\begin{aligned}
C_{n}(y) & =\sum_{v=0}^{\infty} c_{n r} y_{r}=\sum_{r=0}^{\infty}\left(\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{l}_{v r}\right) y_{r} \\
& =\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{L}_{v}(y)=\sum_{v=0}^{n} \widetilde{b}_{n v} T_{v}(x) \\
& =\widetilde{B}_{n}(T(x)) .
\end{aligned}
$$

Also, it can be seen that $C=\widetilde{B} \cdot \bar{L}$. So, by combining the above calculations we get $C \in\left(\ell_{k}, \ell_{s}\right)$. On the other hand, since $\ell_{k}$ is $B K$ space for $k \geq 1$, then, by

Lemma $A$, the matrix $C$ defines a bounded linear operator $L_{C}: \ell_{k} \rightarrow \ell_{s}$ such that $L_{C}(x)=\left(C_{n}(x)\right)$ for all $x \in \ell_{k}$, and so there exists a constant $M$ such that

$$
\begin{equation*}
\left\|L_{C}(x)\right\|_{\ell_{s}} \leq M\|x\|_{\ell_{k}} \text { for all } x \in \ell_{k} \tag{2.7}
\end{equation*}
$$

Now in particular we put $x_{m}=1$ and $x_{n}=0$ for $n \neq m$. Then, we obtain

$$
C_{n}(x)= \begin{cases}0, & n<m \\ c_{n m}, & n \geq m\end{cases}
$$

and

$$
\left\|L_{C}(x)\right\|_{\ell_{s}}=\left(\sum_{n=m}^{\infty}\left|\phi_{n}^{1 / s^{*}} \gamma_{m}^{-1 / k^{*}} \sum_{v=0}^{n} \sum_{i=m}^{\infty} \widehat{b}_{n v} t_{v i} \widehat{a}_{i m}^{\prime}\right|^{s}\right)^{1 / s} .
$$

So, it follows from (2.7) that (2.3) holds. This completes the proof.
Proof of the Theorem 2.2. Let $1=k \leq s<\infty$. Then, $T \in\left(|A|,|\phi B|_{s}\right)$ iff $T(x)$ exists and $T(x) \in|\phi B|_{s}$ for all $x \in|A|$. Now, $x \in|A|$ iff $y \in \ell$, where $y_{n}=\widehat{A}_{n}(x)$ and $\widehat{A}_{n}(x)$ is defined by (1.2). Then, by the inverse of (1.2), we have

$$
x_{n}=\sum_{r=0}^{n} \widehat{a}_{n r}^{\prime} \widehat{A}_{r}(x)=\sum_{r=0}^{n} \widehat{a}_{n r}^{\prime} y_{r},
$$

and so

$$
\begin{aligned}
\sum_{v=0}^{m} t_{n v} x_{v} & =\sum_{v=0}^{m} t_{n v} \sum_{r=0}^{v} \widehat{a}_{v r}^{\prime} \gamma_{r}^{-1 / k^{*}} y_{r} \\
& =\sum_{r=0}^{m}\left(\sum_{v=r}^{m} t_{n v} \widehat{a}_{v r}^{\prime}\right) y_{r}=\sum_{r=0}^{\infty} l_{m r}^{(n)} y_{r} \\
& =L_{m}^{(n)}(y)
\end{aligned}
$$

where

$$
l_{m r}^{(n)}=\left\{\begin{array}{cc}
\sum_{v=r}^{m} t_{n v} \widehat{a}_{v r}^{\prime}, & 0 \leq r \leq m \\
0, & r>m .
\end{array}\right.
$$

This implies that $T(x)$ exists for all $x \in|A|$ iff $L^{(n)}(y) \in(\ell, c)$, or equivalently, by Lemma $B$, (2.4) and (2.5) are satisfied. Further, we have

$$
\begin{aligned}
T_{n}(x) & =\sum_{v=0}^{\infty} t_{n v} x_{v}=\sum_{r=0}^{\infty} \lim _{m \rightarrow \infty} l_{m r}^{(n)} y_{r} \\
& =\sum_{r=0}^{\infty} \bar{l}_{n r} y_{r}=\bar{L}_{n}(y),
\end{aligned}
$$

which also means $T(x)=\bar{L}(y)$. On the other hand, since $T(x)=\bar{L}(y)$, then, $T(x) \in|\phi B|_{s}$ iff $C(y) \in \ell_{s}$, where

$$
c_{n r}=\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{l}_{v r} \text { for } n, r \geq 0
$$

because,

$$
\begin{aligned}
C_{n}(y) & =\sum_{r=0}^{\infty} c_{n r} y_{r}=\sum_{r=0}^{\infty}\left(\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{l}_{v r}\right) y_{r} \\
& =\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{L}_{v}(y)=\sum_{v=0}^{n} \widetilde{b}_{n v} T_{v}(x) \\
& =\widetilde{B}_{n}(T(x)) .
\end{aligned}
$$

Thus it follows from Lemma $C$ that

$$
\sum_{n=0}^{\infty}\left|\sum_{v=0}^{n} \widetilde{b}_{n v} \bar{l}_{v r}\right|^{s}=O(1)
$$

which completes the proof.
We note that in the special case $T=I$, identity matrix, then $I \in\left(|\gamma A|_{k},|\phi B|_{s}\right)$ means that if a series is summable $\left|A, \gamma_{n}\right|_{k}$, then it is also summable $\left|B, \phi_{n}\right|_{s}$, and also, conditions (2.1), (2.2) hold and (2.3) reduces to

$$
\phi_{m}^{s-1}\left|\frac{b_{m m}}{a_{m m}}\right|^{s}+\sum_{n=m+1}^{\infty} \phi_{n}^{s-1}\left|\sum_{i=m}^{n} \widehat{b}_{n i} \widehat{a}_{i m}^{\prime}\right|^{s}=O\left(\gamma_{m}^{s / k^{*}}\right) .
$$

So, as consequences of Theorem 2.1-2.2, we have many results. Now we list some of them.

Corollary 2.3. Let $A$ and $B$ be infinite triangle matrix of complex numbers. Further, let $\left(\gamma_{n}\right)$ and $\left(\phi_{n}\right)$ be two sequences of positive numbers.
a-) If $1<k \leq s<\infty$, then, the necessary conditions in order that a series by summable $\left|A, \gamma_{n}\right|_{k}$ is also summable $\left|B, \phi_{n}\right|_{s}$ are

$$
\begin{gather*}
\phi_{m}^{1 / s^{*}}\left|\frac{b_{m m}}{a_{m m}}\right|=O\left(\gamma_{m}^{1 / k^{*}}\right)  \tag{2.8}\\
\sum_{n=m+1}^{\infty} \phi_{n}^{s-1}\left|\sum_{i=m}^{n} \widehat{b}_{n i} \widehat{a}_{i m}^{\prime}\right|^{s}=O\left(\gamma_{m}^{s / k^{*}}\right) .
\end{gather*}
$$

b-) If $1=k \leq s<\infty$, then, the necessary and sufficient conditions in order that a series by summable $|A|$ is also summable $\left|B, \phi_{n}\right|_{s}$ are that (2.8) and (2.9) with $k=1$ are satisfied.

Let us take $\phi_{n}=\gamma_{n}=n$ for all $n$. Since $\left|A, \gamma_{n}\right|_{k}=|A|_{k}$ and $\left|B, \phi_{n}\right|_{s}=$ $|B|_{s}$, then, Corollary 2.3 reduces to the following result which do not include the additional condition (1.7) of Theorem 1.2.

Corollary 2.4. Let $1<k \leq s<\infty, A$ and $B$ be triangle matrix of complex numbers. Then necessary conditions in order that a series by summable $|A|_{k}$ is also summable $|B|_{s}$ are

$$
m^{1 / k-1 / s}\left|\frac{b_{m m}}{a_{m m}}\right|=O(1)
$$

and

$$
\sum_{n=m+1}^{\infty} n^{s-1}\left|\sum_{i=m}^{n} \widehat{b}_{n i} \widehat{a}_{i m}^{\prime}\right|^{s}=O\left(m^{s / k^{*}}\right)
$$

If $1=k \leq s<\infty$, by Theorem 2.2, these conditions with $k=1$ are also necessary and sufficient for the conclusion to satisfy.

Also, if we put $A=I$ and $\gamma_{v}=v$ for all $v \geq 1$, then the summability $\left|A, \gamma_{n}\right|_{k}$ is equivalent to the condition

$$
\sum_{n=1}^{\infty} n^{k-1}\left|x_{n}\right|^{k}<\infty
$$

Hence the following result is deduced by theorem 2.1, which is due to Sarıgöl [9].
Corollary 2.5. Let $1 \leq s<\infty$ and $B$ be triangle matrix of complex numbers. Then, the necessary and sufficient conditions in order that an absolutely convergent series is also summable $|B|_{s}$ are

$$
\sum_{n=v}^{\infty} n^{s-1}\left|\widehat{b}_{n v}\right|^{s}=O(1)
$$

Further, if $A$ and $B$ are the matrix of weighted means $\left(R, p_{n}\right)$ and $\left(R, q_{n}\right)$ then, it is easily seen that $\widehat{a}_{n v}=p_{n} P_{v-1} / P_{n} P_{n-1}, 1 \leq v \leq n$, and zero otherwise, $\widehat{a}_{v v}^{\prime}=$ $P_{v} / p_{v}, \widehat{a}_{v, v-1}^{\prime}=-P_{v-2} / p_{v-1}$ and $\widehat{a}_{n, v}^{\prime}=0$ for $n \neq v, v+1$, and also, $\widehat{b}_{n v}=$ $q_{n} Q_{v-1} / Q_{n} Q_{n-1}, 1 \leq v \leq n$, and zero otherwise. So the following result follows immediately from Theorem 2.2, of which sufficiency for the case $\phi_{v}=\gamma_{v}=v$ and $k=s$ is due to Orhan and Sarıgöl [5].

Corollary 2.6. Let $1=k \leq s<\infty$ and $B$ be triangle matrix of complex numbers. Then, necessary and sufficient conditions in order that a series by summable $\left|R, p_{n}\right|$ is also summable $\left|R, q_{n}\right|_{s}$ are

$$
v^{1-1 / s}\left|\frac{P_{v} q_{v}}{p_{v} Q_{v}}\right|=O(1)
$$

and

$$
\left|Q_{v-1} \frac{P_{v}}{p_{v}}-Q_{v} \frac{P_{v-1}}{p_{v}}\right|^{s} \sum_{n=v+1}^{\infty} n^{s-1}\left(\frac{q_{n}}{Q_{n} Q_{n-1}}\right)^{s}=O(1) .
$$

Let $A$ and $B$ be Cesàro matrices $(C, \alpha)$ and $(C, \beta)$, respectively. In this case, it is well known that $\widehat{a}_{n v}=v A_{n-v}^{\alpha-1} / n A_{n}^{\alpha}, \widehat{b}_{n v}=v A_{n-v}^{\beta-1} / n A_{n}^{\beta}$, and $\widehat{a}_{n v}^{\prime}=v A_{n-v}^{-\alpha-1} A_{v}^{\alpha} / n$. So, (2.1) is equivalent to

$$
v^{\alpha-\beta+1 / k-1 / s}=O(1)
$$

or $\beta \geq \alpha+1 / k-1 / s$. Also, since (see, Lemma 5, Mehdi [4])

$$
\sum_{n=v}^{\infty} \frac{1}{n}\left|\frac{A_{n-r}^{\beta-\alpha-1}}{A_{n}^{\beta}}\right|^{s}=\left\{\begin{array}{c}
O\left(v^{-s \beta-1}\right), s(\beta-\alpha-1)<-1 \\
O\left(v^{-s \beta-1} \log v\right), s(\beta-\alpha-1)=-1 \\
O\left(v^{-s(\alpha+1)}\right), s(\beta-\alpha-1)>-1
\end{array}\right.
$$

we have

$$
\begin{aligned}
E_{v} & =\sum_{n=v}^{\infty} n^{s-1}\left|\sum_{r=v}^{n} \widehat{b}_{n r} \widehat{a}_{r v}^{\prime}\right|^{s}=\left(v A_{v}^{\alpha}\right)^{s} \sum_{n=v}^{\infty} n^{s-1}\left|\frac{1}{n A_{n}^{\beta}} \sum_{r=v}^{n} A_{n-r}^{\beta-1} A_{r-v}^{-\alpha-1}\right|^{s} \\
& =\left(v A_{v}^{\alpha}\right)^{s} \sum_{n=v}^{\infty} \frac{1}{n}\left|\frac{A_{n-v}^{\beta-\alpha-1}}{A_{n}^{\beta}}\right|^{s}=O\left(v^{s-s / k}\right) .
\end{aligned}
$$

In fact, since $\beta \geq \alpha+1 / k-1 / s$, it is clear that $s(\beta-\alpha-1)+s+1-s / k \geq 0$. So, it is easy to see from Mehdi's lemma that (2.8) is satisfied, because, $E_{v}$ is equal to $O(1) v^{-s(\beta-\alpha-1)-1-s+s / k}, O(1) v^{-s(\beta-\alpha-1)-1-s+s / k} \log v$ and $O(1) v^{-s+s / k}$ for $s(\beta-\alpha-1)<-1, s(\beta-\alpha-1)=-1$ and $s(\beta-\alpha-1)>-1$, respectively. So Theorem 2.1 reduces to the following result of which sufficiency was proved by Flett [2].

Corollary 2.7. Let $1<k \leq s<\infty$, and $\alpha>-1$. Then, necessary conditions in order that a series by summable $|C, \alpha|_{k}$ is also summable $|C, \beta|_{s}$ are $\beta \geq \alpha+$ $1 / k-1 / s$.

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Mehmet Ali Sarıgöl
Pamukkale University
Department of Mathematics
Denizli, Turkey
msarigol@pau.edu.tr

# CESÀRO AND STATISTICAL DERIVATIVE 

Fatih Nuray

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Abstract. In this study, we introduce the notions of Cesàro, strongly Cesàro and statistical derivatives for real valued functions. These notions are based on the concepts of Cesàro and statistical convergence of a sequence. Then we establish some relationships between strongly Cesàro derivative and statistical derivative.
Keywords: Cesàro derivative; statistical derivative; Cesàro continuity; real valued functions; convergence of a sequence.

## 1. Introduction

In mathematical analysis, the concepts of limit, continuity and derivative for a function are given respectively. In the literature, the concept of Cesàro limit has been known for many years. Later, Cesàro continuity, statistical limit and statistical continuity concepts were given (see [5]). In [3] strongly sequentially continuous functions were defined and studied. Cesàro derivative and statistical derivative definitions do not appear in the literature. We will introduce the concepts of Cesàro derivative and statistical derivative in this study to fill the gap in the literature.

A sequence $x=\left(x_{k}\right)$ is said to be Cesàro summable to the number $u$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n} x_{k}=u
$$

in this case we write $(C, 1)-\lim x_{n}=u$, strongly Cesàro summable to the number $u$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n}\left|x_{k}-u\right|=0
$$

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in this case we write $[C, 1]-\lim x_{n}=u$, and statistically convergent to the number $u$ if for every $\epsilon>0$,

$$
\lim _{n \rightarrow \infty} \frac{1}{n}\left|\left\{k \leq n: \quad\left|x_{k}-u\right| \geq \epsilon\right\}\right|=0
$$

where the vertical bars indicate the number of elements in the enclosed set, in this case we write $s t-\lim x_{n}=u$.

Let $\left(a_{n}\right)$ and $\left(b_{n}\right)$ be two sequences of real numbers such that $(C, 1)-\lim a_{n}=a$ and $(C, 1)-\lim b_{n}=b$. It is known that

$$
(C, 1)-\lim a_{n} \cdot b_{n}=a . b \quad \text { and } \quad(C, 1)-\lim \left(a_{n}+b_{n}\right)=a+b .
$$

The idea of statistical convergence was introduced by Steinhaus in [13] and Fast in [6] independently and since then has been studied by other authors including $[4,7,11]$ and $[14]$. Recently, the articles [1], [2], [8], [9] and [10] have been published on statistical convergence and its applications.

## 2. Cesàro Derivative

Very basic finite difference formulas approximates the derivative $f^{\prime}(x)$ using a sequence $x_{n}>0$ such that $\lim _{n \rightarrow \infty} x_{n}=0$. Two basic formulas for derivative of a function $f: \mathbb{R} \rightarrow \mathbb{R}$ at a point $x_{0}$ are

$$
\lim _{n \rightarrow \infty} \frac{f\left(x_{0}+x_{n}\right)-f\left(x_{0}\right)}{x_{n}}=f^{\prime}\left(x_{0}\right) \text { and } \lim _{n \rightarrow \infty} \frac{f\left(x_{0}+x_{n}\right)-f\left(x_{0}-x_{n}\right)}{2 x_{n}}=f^{\prime}\left(x_{0}\right)
$$

The first formula is Newton's difference quotient and determines the slope of a secant line of the graph of $f$. The second formula is the symmetric difference quotient and determines the slope of a cord of the graph of $f$. For more detail (see [12]).
With the similar approach we will now define the Cesàro derivative.
Definition 2.1. A function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a Cesàro derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n} \frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}\right)}{x_{k}}=w
$$

holds whenever $x_{n}>0$ and $\lim _{n \rightarrow \infty} x_{n}=0$.
An equivalent definition to the Definition 2.1 as follows:
Definition 2.2. A function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a Cesàro derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n} \frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}-x_{k}\right)}{2 x_{k}}=w
$$

holds whenever $x_{n}>0$ and $\lim _{n \rightarrow \infty} x_{n}=0$.

A function $f: \mathbb{R} \rightarrow \mathbb{R}$ is Cesàro continuous at a point $x_{0}$ if

$$
(C, 1)-\lim f\left(x_{0}+x_{n}\right)=f\left(x_{0}\right)
$$

holds for each sequence $\left(x_{n}\right) \rightarrow 0$.
Theorem 2.1. Let a function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a Cesàro derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ then $f$ is Cesàro continuous at the point $x_{0}$.

Proof. Let $\lim x_{n}=0$. Clearly

$$
f\left(x_{0}+x_{n}\right)-f\left(x_{0}\right)=\frac{f\left(x_{0}+x_{n}\right)-f\left(x_{0}\right)}{x_{n}} x_{n}
$$

holds for each $n \in \mathbb{N}$. Since $\lim x_{n}=0$ implies $(C, 1)-\lim x_{n}=0$, we can write

$$
(C, 1)-\lim \left(f\left(x_{0}+x_{n}\right)-f\left(x_{0}\right)\right)=(C, 1)-\lim \frac{f\left(x_{0}+x_{n}\right)-f\left(x_{0}\right)}{x_{n}}(C, 1)-\lim x_{n}
$$

Hence, from the assumption we have

$$
(C, 1)-\lim f\left(x_{0}+x_{n}\right)=f\left(x_{0}\right)
$$

so $f$ is Cesàro continuous at the point $x_{0}$.

Definition 2.3. A function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a strongly Cesàro derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n}\left|\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}\right)}{x_{k}}-w\right|=0
$$

holds whenever $x_{n}>0$ and $\lim _{n \rightarrow \infty} x_{n}=0$.
An equivalent definition to the Definition 2.3 as follows:
Definition 2.4. A function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a strongly Cesàro derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n}\left|\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}-x_{k}\right)}{2 x_{k}}-w\right|=0
$$

holds whenever $x_{n}>0$ and $\lim _{n \rightarrow \infty} x_{n}=0$.
It is clear from the definitions of Cesàro and strongly Cesàro derivatives that if a function has a strongly Cesàro derivative at point $x_{0}$, it has a Cesàro derivative at that point.

## 3. Statistical Derivative

In this section, we first give the definition of statistical derivative and then we establish some relationships between the strongly Cesàro derivative and statistical derivative.

Definition 3.1. A function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a statistical derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n}\left|\left\{k \leq n:\left|\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}\right)}{x_{k}}-w\right| \geq \epsilon\right\}\right|=0
$$

holds whenever $x_{n}>0$ and $\lim _{n \rightarrow \infty} x_{n}=0$.
An equivalent definition to the Definition 3.1 as follows:
Definition 3.2. A function $f: \mathbb{R} \rightarrow \mathbb{R}$ has a statistical derivative $w \in \mathbb{R}$ at a point $x_{0} \in \mathbb{R}$ if

$$
\lim _{n \rightarrow \infty} \frac{1}{n}\left|\left\{k \leq n:\left|\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}-x_{k}\right)}{2 x_{k}}-w\right| \geq \epsilon\right\}\right|=0
$$

holds whenever $x_{n}>0$ and $\lim _{n \rightarrow \infty} x_{n}=0$.
If a function has derivative it has statistical derivative but converse may not be true.

Theorem 3.1. a) If a function $f: \mathbb{R} \rightarrow \mathbb{R}$ has strongly Cesàro derivative at a point $x_{0} \in \mathbb{R}$ then it has statistical derivative at the point $x_{0}$.
b) If $\left(\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}\right)}{x_{k}}\right)$ is bounded for each $k \in \mathbb{N}$ and $f$ has statistical derivative at a point $x_{0} \in \mathbb{R}$ then $f$ has strongly Cesàro derivative at the point $x_{0}$.

Proof. Let's write $y_{k}$ instead of $\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}\right)}{x_{k}}$ for simplicity.
a) Let $f$ has strongly Cesàro derivative at a point $x_{0} \in \mathbb{R}$. For an arbitrary $\epsilon>0$, we get

$$
\begin{aligned}
\frac{1}{n} \sum_{k=1}^{n}\left|y_{k}-w\right| & =\left(\frac{1}{n} \sum_{k=1}^{n}\left|y_{k}-w\right|+\frac{1}{n} \sum_{k=1}^{n}\left|y_{k}-w\right| \geq \epsilon\right. \\
& \left.\geq \frac{1}{n} \sum_{k=1\left|y_{k}-w\right|<\epsilon}^{n}\left|y_{k}-w\right|\right) \\
& \left.\geq \frac{1}{n} \right\rvert\,\left\{1 \leq k \leq n:\left|y_{k}-w\right|\right. \\
& \left.\geq y_{k}-w \mid \geq \epsilon\right\} \mid \epsilon
\end{aligned}
$$

Hence, we have

$$
\lim _{n \rightarrow \infty} \frac{1}{n}\left|\left\{1 \leq k \leq n:\left|y_{k}-w\right| \geq \epsilon\right\}\right|=0
$$

that is, $f$ has a statistical derivative at the point $x_{0}$.
b) Now suppose that $f$ has a statistical derivative at the point $x_{0}$ and bounded, since $\left(\frac{f\left(x_{0}+x_{k}\right)-f\left(x_{0}\right)}{x_{k}}\right)$ is bounded for each $k \in \mathbb{N}$, say $\left|y_{k}-w\right| \leq K$ for all $k$. Given $\epsilon>0$, we get

$$
\begin{aligned}
\frac{1}{n} \sum_{k=1}^{n}\left|y_{k}-w\right| & =\frac{1}{n}\left(\sum_{k=1\left|y_{k}-w\right| \geq \epsilon}^{n}\left|y_{k}-w\right|+\sum_{k=1\left|y_{k}-w\right|<\epsilon}^{n}\left|y_{k}-w\right|\right) \\
& \leq \frac{1}{n}\left(K \sum_{k=1}^{n} 1+\sum_{k=1 y_{k}-w \mid \geq \epsilon}^{n}\left|y_{k}-w\right|\right) \\
& \leq K \frac{1}{n}\left|\left\{1 \leq k \leq n: \quad\left|y_{k}-w\right| \geq \epsilon\right\}\right|+\frac{1}{n} \sum_{k=1}^{n} \epsilon
\end{aligned}
$$

hence we have,

$$
\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n}\left|y_{k}-w\right|=0
$$

that is $f$ has strongly Cesàro derivative at the point $x_{0}$.

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Fatih Nuray
Faculty of Science
Department of Mathematics
Afyon Kocatepe University
Afyonkarahisar, Turkey
fnuray@aku.edu.tr

# A HYBRID ALGORITHM FOR THE UNCERTAIN INVERSE $p$-MEDIAN LOCATION PROBLEM 

Akram Soltanpour, Fahimeh Baroughi and Behrooz Alizadeh

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Abstract. In this paper, we investigate the inverse $p$-median location problem with variable edge lengths and variable vertex weights on networks in which the vertex weights and modification costs are the independent uncertain variables. We propose a model for the uncertain inverse $p$-median location problem with tail value at risk objective. Then, we show that it is NP-hard. Therefore, a hybrid particle swarm optimization algorithm has been presented to obtain the approximate optimal solution of the proposed model. The algorithm contains expected value simulation and tail value at risk simulation.
Keywords: $p$-median location problem; inverse optimization; Hybrid algorithm; nonlinear programming.

## 1. Introduction

One of the important aspects of location problems which has recently been studied by many researchers is the $p$-median location problem which can be stated as follows. Let $N=(V, E)$ be an undirected connected network with vertex set $V,|V|=n$, and edge set $E,|E|=m$. The distance between two points on $N$ is equal to the length of the shortest path connecting these two points. Each vertex is associated with a nonnegative weight that is the demand of the client at this vertex. In a $p$-median problem on a network, the aim is to find $p$ locations for establishing facilities on edges or vertices of the network such that the sum of the weighted distances from the clients to the closest facility becomes minimum. In the context of the $p$-median location problems, the interested reader is referred to papers $[1,7,8,11,12,13,17,20,21,28,37,39]$.

In recent years, inverse location problems have found an increasing interest. In an inverse location problem the goal is to modify parameters of the problem at

[^5]minimum cost such that a prespecified solution becomes optimal. Burkard et al. investigated the inverse 1-median problem with variable vertex weights on a tree network and also on a plane and presented algorithms in $O(n \log n)$ time for them [9]. Also they proposed an algorithm in $O\left(n^{2}\right)$ time for the problem under investigation on cycles [10]. Baroughi et al. [3] proved that the inverse $p$-median location problem ( $\mathrm{I} p \mathrm{MLP}$ ) on general networks is NP-hard. For a survey on the inverse $p$-median location problems, we refer the interested reader to $[16,18,19,23,30,36]$.

In the real life, we are usually faced with various types of uncertainty. For example, in location problems, we are usually not sure of the vertex weights, the travel times between vertices, the establishing costs of facilities and the vertex weights or edge lengths modification costs of a network. The uncertainty theory that proposed by Liu [25] is a suitable tool to deal with these parameters. Some researchers applied the uncertainty theory to deal with the location problems,for example Gao [14] modeled the single facility location problems with uncertain demands. Wen et al. [43] investigated the capacitated facility location-allocation problem with uncertain demands and also Nguyen and Chi [31] studied inverse 1-median problem on a tree with uncertain costs and showed that the inverse distribution function of the minimum cost can be obtained at $O\left(n^{2} \log n\right)$ time. For a survey on uncertain location problems, we refer the interested reader to [15, 22, 27, 34, 40, 46].

The uncertainty leads to the risk. Liu in [26] introduced the risk concept in the uncertain environment. Measuring the risk is one of the important steps in the decision making process. The risk metrics contain techniques and data sets used to calculate the risk value of the problem under investigation. Tail value at risk (TVaR) metric [32] is one of the measures of the risk that is widely acceptable among industry segments and market participants.

In the risk management related to location problems, Berman et al. [6] studied the effect of a decision maker's risk attitude on the median and center location problems, with uncertain demand in the mean-variance framework. Wang et al. [41] investigated a two-stage fuzzy facility location problem with value at risk. Wagner et al. [42] developed and examined a new algorithm for solving the p-median problem when the demands are probabilistic and correlated. For a survey on the risk management in the location problems with fuzzy variables, see, e.g. [5, 44].

In this paper, we concentrate on $\mathrm{I} p \mathrm{MLP}$ with variable edge lengths and variable vertex weights on networks. We assume that the vertex weights and modification costs are the independent uncertain variables. We propose a model for the uncertain inverse $p$-median location problem with tail value at risk objective and expected value constraints and show that the problem is NP-hard. Considering the uncertain and NP-hard nature in uncertain I $p$ MLP (UI $p$ MLP), evolutionary and meta-heuristics algorithms can be used to UIpMLP for successful generation of optimal solutions. Hence, we present a hybrid particle swarm optimization algorithm which contains expected value simulation and tail value at risk simulation to obtain the approximate optimal solution of the proposed model.

Based on our knowledge, there are two papers on the implementation of metaheuristic algorithms to the inverse location problems until now. Alizadeh and

Bakhteh [2] studied the general IpMLPs on networks and presented a modified firefly algorithm for the problem under investigation. Mirzapolis Adeh et al. [29] investigated the general inverse ordered $p$-median location problem on crisp networks and designed a modified particle swarm optimization (PSO) algorithm for it. There is no scientific paper on implementation of hybrid metaheuristic algorithms on IpMLPs in uncertain networks. However, many papers can be found in the literature for other classical location problems on uncertain networks. Bashiri et al. [4] modeled fuzzy capacitated $p$-hub center problem and presented a genetic algorithm for the problem. Huang and Hao [22] modeled uncapacitated facility location problem with uncertain customers positions and provided a hybrid intelligent algorithm for solving it. In 2018 Rahmaniani et al. [35] proposed an efficient hybrid solution algorithm for the capacitated facility location-allocation problem under uncertainty. Yang et al. [45] presented an improved hybrid particle swarm optimization algorithm for fuzzy $p$-hub center problem.

The article is organized as follows: In the next section, we first introduce uncertainty theory and TVaR metric in an uncertain environment. Then, we discuss uncertain optimization model and present a new model with TVaR objective and expected value constraints. In Section 3., we first introduce I $p$ MLP with variable edge lengths and variable vertex weights on networks and then investigate the problem with uncertain vertex weights and uncertain modification costs. A model for the uncertain inverse $p$-median location problem (UIpMLP) with TVaR objective is presented and it is shown that the problem under investigation is NP-hard. Then, we present a hybrid PSO algorithm to obtain the approximate optimal solution of the proposed model, which it contains expected value simulation and TVaR simulation. Finally, to show the effectiveness of the proposed hybrid PSO algorithm, we give a numerical example. Section 4. gives a brief conclusion to this paper.

## 2. Preliminaries

In this section, we first present some definitions and theorems of the uncertainty theory and TVaR metric in an uncertain environment. Then, we introduce the uncertain optimization model and present a new model with TVaR objective and expected value constraints.

### 2.1. Uncertainty theory

Let $\Gamma$ be a nonempty set and $\Theta$ be a $\sigma$-algebra over $\Gamma$. An uncertain measure is a set function $\mathcal{M}: \Theta \rightarrow[0,1]$ that satisfies in normality, duality and subadditivity axioms. The triple $(\Gamma, \Theta, \mathcal{M})$ is called an uncertainty space.

Definition 2.1. ( $\operatorname{Liu}[25])$. Let $(\Gamma, \Theta, \mathcal{M})$ be an uncertainty space. A measurable function $\theta$ from $(\Gamma, \Theta, \mathcal{M})$ to the set of real numbers is called an uncertain variable.

Definition 2.2. (Liu[25]). Let $\theta$ be an uncertain variable. For any real number $x$, the function $\Upsilon(x)=\mathcal{M}\{\theta \leq x\}$ is called an uncertainty distribution of $\theta$.

Definition 2.3. (Liu[25]). Let $\theta_{i}, i=1, \ldots, n$, be the uncertain variables. We call $\theta_{i}, i=1, \ldots, n$, independent if for any Borel sets $B_{1}, B_{2}, \ldots, B_{n}$ of real numbers,

$$
\mathcal{M}\left\{\bigcap_{i=1}^{n}\left\{\theta_{i} \in B_{i}\right\}\right\}=\bigwedge_{i=1}^{n} \mathcal{M}\left\{\theta_{i} \in B_{i}\right\}
$$

Definition 2.4. (Liu[25]). The expected value of the uncertain variable $\theta$ is defined as

$$
E[\theta]=\int_{0}^{+\infty} \mathcal{M}\{\theta \geq r\} d r-\int_{-\infty}^{0} \mathcal{M}\{\theta \leq r\} d r
$$

provided that at least one of the two integral is finite.
A real valued function $f\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ is said to be strictly increasing if $f\left(x_{1}, x_{2}, \ldots, x_{n}\right)>f\left(y_{1}, y_{2}, \ldots, y_{n}\right)$ when $x_{i}>y_{i}$ for $i=1,2, \ldots, n$.

Theorem 2.1. (Liu[25]). Let $\theta_{i}, i=1,2, \ldots, n$, be the independent uncertain variables and $\Upsilon_{i}^{-1}, i=1,2, \ldots, n$, be the inverse uncertainty distributions of $\theta_{i}$. Also, let $f\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ be a strictly increasing function with respect to $x_{i}, i=$ $1,2, \ldots, n$. Then the uncertain variable $\vartheta=f\left(\theta_{1}, \theta_{2}, \ldots, \theta_{n}\right)$ has the following inverse uncertainty distribution

$$
\Upsilon^{-1}(\alpha)=f\left(\Upsilon_{1}^{-1}(\alpha), \ldots, \Upsilon_{n}^{-1}(\alpha)\right)
$$

and also it has the following expected value

$$
E[\vartheta]=\int_{0}^{1} f\left(\Upsilon_{1}^{-1}(\alpha), \ldots, \Upsilon_{n}^{-1}(\alpha)\right) d \alpha
$$

### 2.2. TVaR metric in an uncertain environment

The risk demonstrates a situation, in which there is a chance of loss or danger. The quantification of the risk is a key step towards the management and mitigation of the risk. In this section, we introduce the definition of the TVaR metric to account the probability of loss and the severity of the loss in an uncertain environment [32].

In order to define the TVaR metric, we first introduce the definition of the loss function.

Definition 2.5. (Liu[26]). Consider $\theta_{i}, i=1,2, \ldots, n$, as the uncertain factors of a system. A function $f$ is said to be a loss function if some specified loss occurs if and only if

$$
f\left(\theta_{1}, \theta_{2}, \ldots, \theta_{n}\right)>0
$$

In the uncertain environment, TVaR of the loss function is defined as follows.
Definition 2.6. (Peng[32]). Let $\theta_{i}, i=1,2, \ldots, n$, be the uncertain factors and $f$ be the loss function of a system. Then TVaR of $f$ is defined as

$$
T V a R_{\beta}=\frac{1}{\beta} \int_{0}^{\beta} \sup \left\{\lambda \mid \mathcal{M}\left\{f\left(\theta_{1}, \theta_{2}, \ldots, \theta_{n}\right) \geq \lambda\right\} \geq \gamma\right\} d \gamma
$$

for each given risk confidence level $\beta \in(0,1]$.
Theorem 2.2. (Peng[32]). Let $\theta_{i}, i=1,2, \ldots, n$, be the uncertain factors of $a$ system and $\Upsilon_{i}^{-1}, i=1,2, \ldots, n$, be the inverse uncertainty distributions of $\theta_{i}$. Also assume that the loss function $f\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ is a strictly increasing function with respect to $x_{i}, i=1,2, \ldots, n$. Then, for each risk confidence level $\beta \in(0,1]$, we have

$$
T V a R_{\beta}=\frac{1}{\beta} \int_{0}^{\beta} f\left(\Upsilon_{1}^{-1}(1-\gamma), \Upsilon_{2}^{-1}(1-\gamma), \ldots, \Upsilon_{n}^{-1}(1-\gamma)\right) d \gamma
$$

### 2.3. Uncertainty optimization

Let $x=\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ be a decision vector, and $\theta=\left(\theta_{1}, \theta_{2}, \ldots, \theta_{n}\right)$ be an uncertain vector. Consider the following optimization model.

$$
\begin{array}{cll}
\min & f(x, \theta) & \\
\mathrm{s.t.} & g_{j}(x, \theta) \leq 0 & j=1, \ldots, p  \tag{2.1}\\
& z_{l}(x) \leq 0 & l=1, \ldots, m \\
& x \geq 0
\end{array}
$$

where $f$ and $g_{j}, j=1, \ldots, p$ are uncertain functions and $z_{l}, l=1, \ldots, m$ are crisp functions.

Since the objective function of the model (2.1) involves uncertainty, it cannot be directly optimized. Therefore, by considering $f(x, \theta)$ as a loss function, we minimize its TVaR. In addition, since the uncertain constraints do not define a crisp feasible set, we use the expected value of constraints. Thus, the model (2.1) can be reformulated as

$$
\begin{array}{cll}
\min & T V a R_{\alpha}(f(x, \theta)) & \\
\mathrm{s.t.} & E\left(g_{j}(x, \theta)\right) \leq 0 & j=1, \ldots, p,  \tag{2.2}\\
& z_{l}(x) \leq 0 & l=1, \ldots, m \\
& x \geq 0
\end{array}
$$

According to Theorems 2.1 and 2.2, we can rewrite the problem (2.2) as follows:

$$
\begin{array}{cl}
\min & \frac{1}{\beta} \int_{0}^{\beta} f\left(x, \Upsilon_{1}^{-1}(1-\gamma), \Upsilon_{2}^{-1}(1-\gamma), \ldots, \Upsilon_{n}^{-1}(1-\gamma)\right) d \gamma \\
\text { s.t. } & \int_{0}^{1} g_{j}\left(x, \Upsilon_{1}^{-1}(\alpha), \Upsilon_{2}^{-1}(\alpha), \ldots, \Upsilon_{n}^{-1}(\alpha)\right) d \alpha \leq 0 \quad j=1, \ldots, p,  \tag{2.3}\\
& z_{l}(x) \leq 0 \quad l=1, \ldots, m, \\
& x \geq 0,
\end{array}
$$

where $g_{j}\left(x, \theta_{1}, \theta_{2}, \ldots, \theta_{n}\right)$ is strictly increasing with respect to $\theta_{1}, \theta_{2}, \ldots, \theta_{n}$.

## 3. Problem definition

In this section, we first introduce $\mathrm{I} p \mathrm{MLP}$ with variable edge lengths and variable vertex weights on networks and then investigate the problem with uncertain vertex weights and uncertain modification costs. A model for UIpMLP with TVaR objective and expected value constraints is presented. To solve the proposed model, we present a hybrid PSO algorithm which contains expected value simulation and TVaR simulation.

### 3.1. UI $p \mathbf{M L P}$ on networks

We can express $\mathrm{I} p \mathrm{MLP}$ with variable edge lengths and variable vertex weights as follows: Let $N=(V, E)$ with $|V|=n$ and $|E|=m$ be a connected network. Also let vertex $v \in V$ have a positive weight $w(v)$ and edge $e \in E$ have a positive length $\ell_{e}$. In an I $p$ MLP on networks, a set of vertices $\left\{m_{1}, \ldots, m_{p}\right\}$ is given. The goal is to modify $w(v), v \in V$, and $\ell_{e}, e \in E$, at minimum total cost such that the given set becomes a $p$-median of modified location problem. Let us consider nonnegative $\operatorname{costs} c_{e}^{+}$and $c_{v}^{+}$, if $\ell_{e}$ and $w(v)$ are increased by one unit, respectively. Also we consider nonnegative costs $c_{e}^{-}$and $c_{v}^{-}$, if $\ell_{e}$ and $w(v)$ are decreased by one unit, respectively. Let $p_{e}, q_{e}, p_{v}$ and $q_{v}$ be the amounts by which the edge length $\ell_{e}$ and the vertex weight $w(v)$ are increased and decreased, respectively. We let $p_{e}, q_{e}, p_{v}$ and $q_{v}$ obey the upper bounds $u_{e}^{+}, u_{e}^{-}, u_{v}^{+}, u_{v}^{-}$. In addition, assume that $\mathcal{S}$ is the set of all subsets $S \subseteq V$ with $|S|=p$. Thus, I $p$ MLP on $N$ can be stated as follows.

Change $\ell_{e}, e \in E$, to $\tilde{\ell}_{e}=\ell_{e}+p_{e}-q_{e}$ and $w(v), v \in V$, to $\tilde{w}(v)=w(v)+p_{v}-q_{v}$ such that
(i) The set $\left\{m_{1}, \ldots, m_{p}\right\}$ becomes a $p$-median of $N$ with respect to $\tilde{\ell}$ and $\tilde{w}(v)$, i.e.,

$$
\begin{equation*}
\sum_{v \in V} \tilde{w}(v) \min _{i=1, \ldots, p} d_{\tilde{\ell}}\left(v, m_{i}\right) \leq \sum_{v \in V} \tilde{w}(v) \min _{k \in S} d_{\tilde{\ell}}\left(v, v_{k}\right) \quad \forall S \in \mathcal{S} \tag{3.1}
\end{equation*}
$$

(ii) The bound constraints are satisfied:

$$
\begin{align*}
& 0 \leq p_{e} \leq u_{e}^{+}, \quad 0 \leq q_{e} \leq u_{e}^{-} \quad \forall e \in E  \tag{3.2}\\
& 0 \leq p_{v} \leq u_{v}^{+}, \quad 0 \leq q_{v} \leq u_{v}^{-} \quad \forall v \in V \tag{3.3}
\end{align*}
$$

(iii) The objective function

$$
\sum_{e \in E}\left(c_{e}^{+} p_{e}+c_{e}^{-} q_{e}\right)+\sum_{v \in V}\left(c_{v}^{+} p_{v}+c_{v}^{-} q_{v}\right)
$$

becomes minimum.

This formulation of $\mathrm{I} p \mathrm{MLP}$ is a nonlinear programming model. In the following, we consider I $p$ MLP with uncertain vertex weights and uncertain modification costs.

Let $N=(V, E)$ be a network with independent uncertain vertex weights $\theta_{v}, v \in V$. Also let $w(v)$ be a parameter on each vertex $v \in V$, which will be changed to $\tilde{w}(v)$. In addition, suppose that $\theta_{v}$ relates to this parameter, i.e., for each vertex $v \in V$, we have an original weight $\theta(w(v))$ and also a new weight $\theta(\tilde{w}(v))$. Let $\vartheta_{v}^{+}$and $\vartheta_{v}^{-}$ be the independent uncertain variables with respect to the costs $c_{v}^{+}$and $c_{v}^{-}$, for all $v \in V$, and $\vartheta_{e}^{+}$and $\vartheta_{e}^{-}$be the independent uncertain variables with respect to the $\operatorname{costs} c_{e}^{+}$and $c_{e}^{-}$, for all $e \in E$, respectively.

Let us assume that we are given a set of vertices $\left\{m_{1}, \ldots, m_{p}\right\}$. In an UIpMLP, the goal is to find $\tilde{\ell}_{e}=\ell_{e}+p_{e}-q_{e}$ and $\tilde{w}(v)=w(v)+p_{v}-q_{v}$ such that $\left\{m_{1}, \ldots, m_{p}\right\}$ becomes a $p$-median of the problem with respect to $\theta_{v}(\tilde{w}(v))$ and $\tilde{\ell}_{e}, v \in V, e \in E$, and the total cost

$$
\sum_{v \in V}\left(\vartheta_{v}^{+} p_{v}+\vartheta_{v}^{-} q_{v}\right)+\sum_{e \in E}\left(\vartheta_{e}^{+} p_{e}+\vartheta_{e}^{-} q_{e}\right)
$$

is minimized.
Therefore, we can model UIpMLP as follows.

$$
\min \left[\sum_{v \in V}\left(\vartheta_{v}^{+} p_{v}+\vartheta_{v}^{-} q_{v}\right)+\sum_{e \in E}\left(\vartheta_{e}^{+} p_{e}+\vartheta_{e}^{-} q_{e}\right)\right]
$$

s.t.

$$
\begin{align*}
& {\left[\sum_{v \in V} \theta(\tilde{w}(v))\left(\min _{i=1, \ldots, p} d_{\tilde{\ell}}\left(v, m_{i}\right)-\min _{k \in S} d_{\tilde{\ell}}\left(v, v_{k}\right)\right)\right] \leq 0 \quad \forall S \in \mathcal{S},}  \tag{3.4}\\
& 0 \leq p_{e} \leq u_{e}^{+}, \quad 0 \leq q_{e} \leq u_{e}^{-} \quad \forall e \in E \\
& 0 \leq p_{v} \leq u_{v}^{+}, \quad 0 \leq q_{v} \leq u_{v}^{-} \quad \forall v \in V
\end{align*}
$$

Definition 3.1. Let $p=\left(p_{e}\right)_{e \in E}$ and $q=\left(q_{v}\right)_{v \in V}$ be the vectors that satisfies in (3.2) and (3.3). Then $(p, q)$ is called expected solution of (3.4) if and only if $\forall S \in \mathcal{S}$

$$
\sum_{v \in V} E[\theta(\tilde{w}(v))]\left(\min _{i=1, \ldots, p} d_{\tilde{\ell}}\left(v, m_{i}\right)-\min _{k \in S} d_{\tilde{\ell}}\left(v, v_{k}\right)\right) \leq 0
$$

Now, let $(p, q)$ be a expected solution of (3.4). Define

$$
f(p, q)=\sum_{v \in V}\left(\vartheta_{v}^{+} p_{v}+\vartheta_{v}^{-} q_{v}\right)+\sum_{e \in E}\left(\vartheta_{e}^{+} p_{e}+\vartheta_{e}^{-} q_{e}\right)
$$

Definition 3.2. For a risk confidence level $\beta \in(0,1]$, a expected solution $\left(p^{*}, q^{*}\right)$ is called optimal solution with minimum $T V a R$ if

$$
T V a R_{\beta}\left(f\left(p^{*}, q^{*}\right)\right) \leq T V a R_{\beta}(f(p, q))
$$

holds for any expected solution $(p, q)$.
Therefore, we can find an optimal expected solution with minimum TVaR as follows:

Let $\left(\Psi_{v}^{+}\right)^{-1}, v \in V$, and $\left(\Psi_{e}^{+}\right)^{-1}, e \in E$ be the inverse uncertainty distributions of $\vartheta_{v}^{+}$and $\vartheta_{e}^{+}$, respectively. Also let $\left(\Psi_{v}^{-}\right)^{-1}, v \in V$, and $\left(\Psi_{e}^{-}\right)^{-1}, e \in E$ be the inverse uncertainty distributions of $\vartheta_{v}^{-}$and $\vartheta_{e}^{-}$, respectively. Assume that $\Upsilon_{v}^{-1}$, $v \in V$, is the inverse uncertainty distribution of $\theta_{v}$. Then, for a risk confidence level $\beta \in(0,1]$, the optimal expected solution with minimum TVaR is the optimal solution of the following model:

$$
\begin{array}{ll}
\min & \sum_{v \in V}\left[\left(\frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{v}^{+}\right)^{-1}(1-\gamma) d \gamma\right) p_{v}+\left(\frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{v}^{-}\right)^{-1}(1-\gamma) d \gamma\right) q_{v}\right] \\
& +\sum_{e \in E}\left[\left(\frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{e}^{+}\right)^{-1}(1-\gamma) d \gamma\right) p_{e}+\left(\frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{e}^{-}\right)^{-1}(1-\gamma) d \gamma\right) q_{e}\right] \\
\text { s.t. } &  \tag{3.5}\\
& \sum_{v \in V}\left(\int_{0}^{1} \Upsilon_{v}^{-1}(\tilde{w}(v), \alpha) d \alpha\right)\left(\min _{i=1, \ldots, p} d_{\tilde{\ell}}\left(v, m_{i}\right)-\min _{k \in S} d_{\tilde{\ell}}\left(v, v_{k}\right)\right) \leq 0 \\
& 0 \leq p_{e} \leq u_{e}^{+}, \quad 0 \leq q_{e} \leq u_{e}^{-} \quad \forall e \in E, \\
& 0 \leq p_{v} \leq u_{v}^{+}, \quad 0 \leq q_{v} \leq u_{v}^{-} \quad \forall v \in V
\end{array}
$$

The above model is a deterministic inverse $p$-median problem formulation with vertex weights

$$
\int_{0}^{1} \Upsilon_{v}^{-1}(\tilde{w}(v), \alpha) d \alpha, \quad \forall v \in V
$$

vertex weight modification costs

$$
\frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{v}^{+}\right)^{-1}(1-\gamma) d \gamma, \frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{v}^{-}\right)^{-1}(1-\gamma) d \gamma
$$

and edge length modification costs

$$
\frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{v}^{+}\right)^{-1}(1-\gamma) d \gamma, \frac{1}{\beta} \int_{0}^{\beta}\left(\Psi_{v}^{-}\right)^{-1}(1-\gamma) d \gamma
$$

Baroughi et al. in [3] showed that $\mathrm{I} p \mathrm{MLP}$ on general networks is NP-hard. Thus we immediately conclude the following proposition.

Proposition 3.1. UIpMLP with TVaR criterion on general networks is NP-hard.

The above proposition implies that it is not possible to present exact polynomial time methods to solve UIpMLP on general networks. Therefore, we propose an efficient hybrid PSO algorithm for approximating the optimal solution of UIp MP on networks.

### 3.2. Hybrid PSO algorithm

Kennedy and Eberhart in 1995 [24] developed the PSO algorithm as a natureinspired evolutionary computation algorithm. Consider the following model

$$
\begin{array}{cl}
\min & f(x) \\
\text { s.t. } & x \in X,
\end{array}
$$

where $X$ is the restricted region. In PSO algorithm, a potential solution is presented as a particle $x_{j} \in X$ and a direction $v_{j} \in \mathbb{R}$ in which the particle will move. A swarm of particles is defined as a set $\left\{x_{1}, x_{2}, \ldots, x_{N}\right\}$, in which $N$ is number of particles. Each particle $x_{j}$ retains a record of the position of its previous best performance in a vector called $P_{b e s t, j}$. The particle with best performance r in the population has been maintained in a vector $G_{b e s t}$. An iteration involves evaluating of each particle $x_{j}$, then randomly setting of $v_{j}$ in the direction of particle $x_{j}^{\prime}$ s best previous position $P_{b e s t, j}$ and the best previous position $G_{b e s t}$ of any particle in the population.

Since in UIpMLP the aim is to modify the vertex weights and edge lengths with respect to modification bounds. Thus, we consider a particle of the problem as $x_{j}=\left(x_{1, j}, x_{2, j}, \ldots, x_{2 m+2 n, j}\right)$ where

$$
\begin{align*}
& \left(x_{1, j}, x_{2, j}, \ldots, x_{m, j}\right)=\left(p_{e}\right)_{e \in E} \\
& \left(x_{m+1, j}, x_{m+2, j}, \ldots, x_{2 m, j}\right)=\left(q_{e}\right)_{e \in E} \\
& \left(x_{2 m+1, j}, x_{2 m+2, j}, \ldots, x_{2 m+n, j}\right)=\left(p_{v}\right)_{v \in V}  \tag{3.6}\\
& \left(x_{2 m+n+1, j}, x_{2 m+n+2, j}, \ldots, x_{2 m+2 n, j}\right)=\left(q_{v}\right)_{v \in V} .
\end{align*}
$$

Therefore, $x_{j}$ represents the decision vector of UIpMLP that used in PSO. In addition, according to the orthogonality condition

- if $q_{e}>p_{e}$, then $q_{e}=q_{e}-p_{e}, p_{e}=0$,
- if $q_{e}<p_{e}$, then $p_{e}=p_{e}-q_{e}, q_{e}=0$,
- if $q_{v}>p_{v}$, then $q_{v}=q_{v}-p_{v}, p_{v}=0$,
- if $q_{v}<p_{v}$, then $p_{v}=p_{v}-q_{v}, q_{v}=0$.

For checking the feasibility of particle $x_{j}$, we calculate the expected value of constraints by using the following uncertain simulation algorithm [33]. Let $S \in \mathcal{S}$.

Algorithm 1 (Expected value simulation)

1. Set $E=0$.
2. For $k=1, \ldots, 99$ do
compute

$$
E_{k}=0.01 \sum_{v \in V}\left(\Upsilon_{v}^{-1}(\tilde{w}(v), 0.0 k)\right)\left(\min _{i=1, \ldots, p} d_{\tilde{\ell}}\left(v, m_{i}\right)-\min _{k \in S} d_{\tilde{\ell}}\left(v, v_{k}\right)\right),
$$

and $E:=E+E_{k}$.

## 3. Report $E$.

Therefore, if the particle $x_{j}=\left(x_{1, j}, x_{2, j}, \ldots, x_{2 m+2 n, j}\right)$ is defined as (3.6) and for each $S \in \mathcal{S}, E \leq 0$, then $x_{j}$ is feasibile.

Based on Theorem 2.2, we present the following uncertain simulation procedure for computing TVaR of objective function for each feasible particle $x_{j}$ and given $\beta \in(0,1]$.

Algorithm 2 (TVaR simulation)

1. Set $T_{\beta}=0$.
2. For $j=1, \ldots, M$ do
compute

$$
\begin{aligned}
T_{\beta}^{j} & =\sum_{v \in V}\left[\left(\left(\Psi_{v}^{+}\right)^{-1}\left(1-\frac{j}{M} \beta\right)\right) p_{v}+\left(\left(\Psi_{v}^{-}\right)^{-1}\left(1-\frac{j}{M} \beta\right)\right) q_{v}\right] \\
& +\sum_{e \in E}\left[\left(\left(\Psi_{e}^{+}\right)^{-1}\left(1-\frac{j}{M} \beta\right)\right) p_{e}+\left(\left(\Psi_{e}^{-}\right)^{-1}\left(1-\frac{j}{M} \beta\right)\right) q_{e}\right]
\end{aligned}
$$

and $T_{\beta}=T_{\beta}+\frac{j}{M} \beta T_{\beta}^{j}$.
3. Compute $T V a R_{\beta}=\frac{1}{\beta} T_{\beta}$.

## 4. Report $T V a R_{\beta}$.

To solve the model (3.5) with hybrid PSO algorithm, we first randomly generate the particle $x_{j}$ by checking the feasibility of it using expected value simulation. Repeat this process $N$ times. We get $N$ initial feasible particles $x_{1}, x_{2}, \ldots, x_{N}$. Then, we assume that the fitness of each $x_{j}$ is the minus of TVaR, i.e.,

$$
\operatorname{Fit}\left(x_{j}\right)=-T V a R_{\beta}\left(x_{j}\right)
$$

Thus, the particle with higher fitness has smaller objective value. The fitness of each particle is obtained by using TVaR simulation.

In the process of updating $(i+1)$ th iteration, we first denote $P_{\text {best }, j}(i)$ for each particle $x_{j}$ and $G_{b e s t}(i)$, then we obtain the new directs and the positions of the particles by using the following two equations:

$$
\begin{equation*}
v_{j}(i+1)=v_{j}(i)+C_{1} r_{1}\left[P_{\text {best }, j}(i)-x_{j}(i)\right]+C_{2} r_{2}\left[G_{\text {best }}(i)-x_{j}(i)\right] \tag{3.7}
\end{equation*}
$$

$$
\begin{equation*}
x_{j}(i+1)=x_{j}(i)+v_{j}(i+1) \tag{3.8}
\end{equation*}
$$

where, $P_{\text {best }, j}(i)=x_{j}(i)$ if

$$
\operatorname{Fit}\left(x_{j}(i)\right) \geq \operatorname{Fit}\left(x_{j}(i-1)\right)
$$

and $P_{\text {best }, j}(i)=P_{\text {best }, j}(i-1)$ otherwise, and $G_{\text {best }}(i)=P_{\text {best }, k}(i)$, with

$$
k=\operatorname{argmin}\left\{P_{\text {best }, j}(i): j=1, \ldots, N\right\} .
$$

In addition $r_{1}$ and $r_{2}$ are uniformly distributed random numbers in the interval $[0,1]$ and $C_{1}$ and $C_{2}$ are learning rates, to well adjust the convergence of the particles. The values of $C_{1}$ and $C_{2}$ are usually assumed to be 2 .

If the updated $x_{j}$ is feasible, then we consider it as a new particle of the next generation. Otherwise, as long as a feasible new particle is found, we re-update (3.7) and (3.8).

We obtain a new generation of particles by repeating the above process $N$ times.
If MaxIt indicate the number of generations of the PSO algorithm, then based on all the explanations above, we summarize the hybrid PSO algorithm for solving the model (3.5) as follows.

Algorithm 3 (Hybrid PSO algorithm)

1. Initialize the feasibile particles $x_{1}, \ldots, x_{N}$ (use expected value simulation).
2. Compute the fitness for all particles by using TVaR simulation, and evaluate each particle according to it.
3. Update all the particles by using equations (3.7) and (3.8).
4. As long as a new feasible population is found, re-update (3.7) and (3.8).
5. Repeat Steps 2 to 4 for MaxIt times.
6. Return $G_{b e s t}$ as the optimal solution of the model (3.5), and

$$
T V a R_{\beta}\left(G_{b e s t}\right)=-F i t\left(G_{b e s t}\right)
$$

as the corresponding optimal value.

### 3.3. An illustrative example

In this subsection, we give a numerical example to illustrate the hybrid PSO algorithm. The result of the numerical experiment is obtained on a PC with processor Intel(R) Core(TM) i3 CPU 2.27 GHZ and 4 GB of RAM under windows 7.

We apply the hybrid PSO algorithm for solving UIpMLP with TVaR criteria at a risk confidence level of $\beta=0.8$ on the given network $N$ in Figure 3.1. Let the
cost coefficients be linear uncertain variables (see Table 3.3). Also let the vertex weights $\theta$ be the linear uncertain variables with respect to, $\tilde{w}(v)$, i.e.,

$$
\theta=\theta(\tilde{w}(v))=\mathcal{L}(\tilde{w}(v)-10, \tilde{w}(v)+10)
$$

The input data of the network are given in Tables 3.1 and 3.3.

Note that if $\theta=\mathcal{L}(a, b)$ is the linear uncertain variable, then for a risk confidence level $\beta \in(0,1]$

$$
T V a R_{\beta}(\theta)=\frac{\beta}{2}(a-b)+b
$$

and

$$
E[\theta]=\frac{(a+b)}{2}
$$

In the following, we show the computational results of the hybrid PSO algorithm's performance on an example of UI2MLP on the given network.

Note that the goal is to change $w(v)$ and $\ell_{e}$ with respect to modification bounds so that $\left\{v_{2}, v_{3}\right\}$ becomes a 2 -median at minimum total cost under the new vertex weights and edge lengths.


Fig. 3.1: Network $N$

Table 3.1: The input data for UI2MLP

| $\ell_{e}$ | $(14,34,25,7,22,10,8,20,12,7,10,26,12,6,10,23,31,21,22)$ |
| :---: | :---: |
| $u_{e}^{+}$ | $(5,4,5,4,7,2,5,6,3,9,2,13,1,3,5,1,7,4,1)$ |
| $u_{e}^{-}$ | $(10,30,15,3,17,8,4,10,8,4,5,20,6,1,8,13,2,3,13)$ |
| $w(v)$ | $(34,18,14,13,21,11,13,20,40,22,9,17,13,6,24,14,15,12)$ |
| $u_{v}^{+}$ | $(20,11,4,1,2,4,7,8,15,32,13,5,15,1,2,4,1,1)$ |
| $u_{v}^{-}$ | $(3,2,11,9,10,8,1,1,6,7,2,2,2,4,15,10,13,11)$ |

The hybrid PSO algorithm is run for the problem with $100,200,300$ and 400 generations, respectively. Table 3.2 shows the best solutions of the problem.

Table 3.4, shows the best solutions of UI2MLP using hybrid PSO algorithm. Furthermore, the convergence of the objective values with population sizes $10,15,20,25$ and MaxIt $=100$ is shown in Figure 3.2. The convergence of the objective values with $N=10$ and MaxIt $=100,200,300,400$ is given in Figure 3.3.

Table 3.2: The results of the performance of hybrid PSO algorithm

| N, MaxIt | Objective value | N, MaxIt | Objective value |
| :---: | :---: | :---: | :---: |
| 10,100 | -6880 | 10,200 | -6635 |
| 15,100 | -8240 | 10,300 | -7609 |
| 20,100 | -8520 | 10,400 | -9150 |
| 25,100 | -10496 |  |  |

Table 3.3: Uncertain cost coefficients

| $\vartheta_{e}^{+}$ | $(\mathcal{L}(8,10), \mathcal{L}(18,21), \mathcal{L}(19,21), \mathcal{L}(4,6), \mathcal{L}(3,4), \mathcal{L}(14,16), \mathcal{L}(28,30), \mathcal{L}(10,12)$, |
| :---: | :---: |
|  | $\mathcal{L}(17,18), \mathcal{L}(6,8), \mathcal{L}(22,24), \mathcal{L}(7,9), \mathcal{L}(15,17), \mathcal{L}(18,21), \mathcal{L}(26,28)$, |
|  | $\mathcal{L}(28,30), \mathcal{L}(16,18), \mathcal{L}(4,6), \mathcal{L}(4,6))$ |
| $\vartheta_{e}^{-}$ | $(\mathcal{L}(18,20), \mathcal{L}(14,15), \mathcal{L}(10,12), \mathcal{L}(24,26), \mathcal{L}(17,18), \mathcal{L}(15,17), \mathcal{L}(27,29), \mathcal{L}(8,10)$, |
|  | $\mathcal{L}(22,24), \mathcal{L}(22,24), \mathcal{L}(11,13), \mathcal{L}(2,4), \mathcal{L}(1,3), \mathcal{L}(15,17), \mathcal{L}(24,25)$, |
|  | $\mathcal{L}(28,30), \mathcal{L}(3,5), \mathcal{L}(17,19), \mathcal{L}(17,19))$ |
| $\vartheta_{v}^{+}$ | $(\mathcal{L}(24,26), \mathcal{L}(27,28), \mathcal{L}(3,5), \mathcal{L}(27,28), \mathcal{L}(19,20), \mathcal{L}(1,4), \mathcal{L}(8,10), \mathcal{L}(16,18)$, |
|  | $\mathcal{L}(29,30), \mathcal{L}(29,30), \mathcal{L}(4,6), \mathcal{L}(30,31), \mathcal{L}(29,30), \mathcal{L}(14,16), \mathcal{L}(24,26)$, |
|  | $\mathcal{L}(4,6), \mathcal{L}(12,13), \mathcal{L}(27,28))$ |
| $\vartheta_{v}^{--}$ | $(\mathcal{L}(19,21), \mathcal{L}(22,24), \mathcal{L}(12,13), \mathcal{L}(19,21), \mathcal{L}(5,6), \mathcal{L}(20,22), \mathcal{L}(1,2), \mathcal{L}(7,10)$, |
|  | $\mathcal{L}(2,3), \mathcal{L}(3,4), \mathcal{L}(24,26), \mathcal{L}(20,21), \mathcal{L}(19,20), \mathcal{L}(29,30), \mathcal{L}(2,3)$, |
|  | $\mathcal{L}(13,15), \mathcal{L}(12,13), \mathcal{L}(22,24))$ |

Table 3.4: The obtained $G_{b e s t}$ for UI2MLP by using hybrid PSO algorithm

| N, MaxIt | $G_{\text {best }}$ |
| :---: | :---: |
| 10, 100 | $(3.92,0.90,4.88,0,4.84,0,0,5.82,1.25,0,0,0.33,0,0,0.99,0.94$, $0,0.82,0.69,0,0,0,2.75,0,2.50,3.22,0,0,2.80,1.08,0,4.60,0.23,0$, $0,1.46,0,0,0,9.51,3.40,0.32,1.38,0,0,0,0,0,0,4.94,0,0,1.54$, $3.26,0.61,0,0.22,0,0,0,0,6.31,0.34,0.72,4.28,5.64,0.01,0,0.34,2.56$, $0,0,0,10.73)$ |
| 15, 100 | $(0.24,2.85,4.07,0,4.07,1.29,1.95,4.12,0,8.92,0,8.38,0.97,0.30$, $3.2,0,3.42,3.49,0.08,0,0,0,2.12,0,0,0,0,6.16,0,2.68,0,0,0,0$, $7.84,0,0,0,0.50,0,1.95,0.30,0.94,3.70,6.81,0,0,7.51,5.21,1.95$, $0,0,0,0,0,0,0,0.94,0,0,0,0,0,0.31,0.48,0,0,0,0.8,1.08,4.82$, $8.40,5.80,10.03)$ |
| 20, 100 | $(3.43,3.91,1.20,0,4.02,0,0,4,2.57,0,1.48,11.32,0,0,3.03,0,0,0$, $0,0,0,0,1.67,0,3.10,1.55,0,0,3.70,0,0,0.86,0.96,0,7.83,1.09$, $1.22,2.76,0.52,0,0.78,0,0.44,1.79,6.47,0,11.03,0,0,0,4.89,0.02,0$, $0,0.29,0,0,0.11,0,4.76,0,0,0,0.10,0,5.28,1.77,0.68,0,0,4.57,5.31$, $0,10)$ |
| 25, 100 | $(2.32,2.36,2.88,0,4.76,0,0.28,0,0.80,0,1.78,0,0.16,0,2.61$, $0,1.70,0,0.91,0,0,0,2.04,0,1.45,0,9.33,0,0.50,0,8.00,0,0.39,0$, $7.15,0,1.06,0,0,8.81,0.69,0,0,0.93,1.37,4.04,3.59,26.46,0,0.82,0$ $0,0,0,0.45,0,0.62,0,0,4.63,4.86,0,0,0,0,0,1.35,0,0.12,0.72$, $12.62,3.096,0,10.65)$ |
| 10, 200 | (3.25, 2.75, 3.77, 0, 5.40, 1.71, 2.35, 0, 2.47, 4.18, 0, 8.17, 0, 0, 4.96, 0 , <br> $0,0,0.56,0,0,0,1.39,0,0,0,9.88,0,0,4.93,0,4.93,0.29,0,3.01$, <br> $1.57,1.16,0,0,1.42,0,0.85,0,3.90,2.67,5.46,0,0,0,0.95,0.58,0.72$, $0,0,0.53,0,2.24,0,4.77,0,4.97,0,0,0,5.12,6.30,1.55,0,0,0,9.99$, <br> $1.85,0,10.17$ ) |
| 10, 300 | $4.42,0.91,2.20,0,2.82,0,0,0,2.54,0,0,1.42,0,0,0,0,4.70,0,0.77$, $0,0,0,1.51,0,1.05,1.30,5.94,0,0.89,2.83,0,2.91,0.20,1.74,2.71$, $0,2.60,0,0,3.66,0,0,0,0,6.34,6.11,9.43,0,0,1.10,0,0.20,1.64$, $0.36,0,0,2.31,0,1.27,2.86,6.34,5.40,0,0,0,4.33,0.79,0,0.50,0$, $0,0,0.81,10.66)$ |
| 10, 400 | $\begin{gathered} \text { (1.80, 2.94, 1.64, 0, 3.60, 0, 2.13, 4.58, 0.53, 0, 0.36, 11.20, 0.42, } 1.08, \\ 0,0,0,2.98,0,0,0,0,1.94,0,2.28,0,0,0,3.89,0,0,0,0,7.56,5.33, \\ 1.80,0,9.03,1.69,8.02,1.17,0,0,3.97,1.14,7.81,0,1.61,4.24, \\ 1.71,0,0.71,0.71,0,0.66,0,0,0,0,8.40,3.35,0,0,0,4.31,0,0,0, \\ 1.43,0,0,4.72,0,10.97) \\ \hline \end{gathered}$ |



Fig. 3.2: The convergence of TVaR, MaxIt=100


Fig. 3.3: The convergence of TVaR, $\mathrm{N}=10$

## 4. Conclusion

In this paper, we investigated $\mathrm{I} p \mathrm{MLP}$ with variable edge lengths and variable vertex weights on a network in which the vertex weights and modification costs are the independent uncertain variables. We proposed a model for UIpMLP with TVaR objective and expected value constraints and showed that it is NP-hard. Thus, we presented a hybrid PSO algorithm for approximating the optimal solutions, which it contains expected value simulation and TVaR simulation. Finally, by computational experiments, the efficiency of the algorithm is illustrated.

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Akram Soltanpour
Faculty of Basic Sciences
Department of Applied Mathematics
Sahand University of Technology, Tabriz, Iran

Fahimeh Baroughi
Faculty of Basic Sciences
Department of Applied Mathematics
Sahand University of Technology, Tabriz, Iran
baroughi@sut.ac.ir
(Corresponding author)

Behrooz Alizadeh
Faculty of Basic Sciences
Department of Applied Mathematics
Sahand University of Technology, Tabriz, Iran

# MULTIPLE USE OF BACKTRACKING LINE SEARCH IN UNCONSTRAINED OPTIMIZATION 

Branislav Ivanov, Bilall I. Shaini and Predrag S. Stanimirović

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#### Abstract

The class of gradient methods is a very efficient iterative technique for solving unconstrained optimization problems. Motivated by recent modifications of some variants of the SM method, this study proposed two methods that are globally convergent as well as computationally efficient. Each of the methods is globally convergent under the influence of a backtracking line search. Results obtained from the numerical implementation of these methods and performance profiling show that the methods are very competitive with respect to well-known traditional methods.


Keywords: unconstrained optimization; gradient methods; line search.

## 1. Introduction

The following unconstrained optimization problem

$$
\begin{equation*}
\min _{\mathbf{x} \in \mathbb{R}^{n}} f(\mathbf{x}) \tag{1.1}
\end{equation*}
$$

is ubiquitous in all areas of science and practical engineering applications. In (1.1), the function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is uniformly convex (UC) and twice continuously differentiable (TCD).

The most frequent iterations for solving (1.1) is the gradient descent (GD) iterative scheme

$$
\begin{equation*}
\mathbf{x}_{k+1}^{G D}=\mathbf{x}_{k}^{G D}-t_{k} \mathbf{g}_{k} \tag{1.2}
\end{equation*}
$$

where $t_{k}>0$ is the stepsize and $\mathbf{g}_{k}:=\nabla f\left(\mathbf{x}_{k}\right)$ corresponds to the gradient of $f$. The step length $t_{k}$ is mainly calculated using the backtracking line search (BLS).

The Newton iterations stabilized by the line search are defined as

$$
\begin{equation*}
\mathbf{x}_{k+1}=\mathbf{x}_{k}-t_{k} G_{k}^{-1} \mathbf{g}_{k} \tag{1.3}
\end{equation*}
$$

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wherein $G_{k}^{-1}$ means the inverse of the Hessian matrix $G_{k}:=\nabla^{2} f\left(\mathbf{x}_{k}\right)$. In order to avoid time consuming computation of the Hessian and its inverse, practical numerical methods for solving unconstrained optimization problem are derived from the usage of appropriate approximations $H_{k}$ of $G_{k}^{-1}$. The general scheme of quasiNewton type with line search [16] is given by

$$
\begin{equation*}
\mathbf{x}_{k+1}=\mathbf{x}_{k}-t_{k} H_{k} \mathbf{g}_{k} . \tag{1.4}
\end{equation*}
$$

In order to define efficient class of quasi-Newton methods, we use the simplest scalar approximation of the Hessian with respect to known classifications from [5, 8]:

$$
\begin{equation*}
B_{k}:=\gamma_{k} I, \gamma_{k}>0, \tag{1.5}
\end{equation*}
$$

where $I$ is an identity matrix of appropriate order and $\gamma_{k}>0$ is a real parameter. The choice (1.5) leads to the iterative prototype

$$
\begin{equation*}
\mathbf{x}_{k+1}=\mathbf{x}_{k}-\gamma_{k}^{-1} t_{k} \mathbf{g}_{k} \tag{1.6}
\end{equation*}
$$

where $t_{k}$ denotes the basic step size and $\gamma_{k}^{-1}$ is an additional step size which should be defined appropriately. Clearly, the value $\gamma_{k}^{-1} t_{k}$ can be considered as a composite step size, so that iterations (1.6) are GD methods. The iterations (1.6) are known as improved gradient descent (IGD) methods.

Andrei in $[1,3]$ originated so called Accelerated Gradient Descent (AGD) iterations in the form

$$
\begin{equation*}
\mathbf{x}_{k+1}^{A G D}=\mathbf{x}_{k}^{A G D}-\theta_{k}^{A G D} t_{k} \mathbf{g}_{k} \tag{1.7}
\end{equation*}
$$

The AGD process (1.7) was improved into the Modified AGD (MAGD) method [7] as

$$
\begin{equation*}
\mathbf{x}_{k+1}^{M A G D}=\mathbf{x}_{k}^{M A G D}-\theta_{k}\left(t_{k}+t_{k}^{2}-t_{k}^{3}\right) \mathbf{g}_{k} \tag{1.8}
\end{equation*}
$$

A few variants of the IGD class (1.6) were proposed in [7, 10, 11, 14, 15]. The $S M$ method belongs to the class IGD methods. It was originated in [14] by the iterative process

$$
\begin{equation*}
\mathbf{x}_{k+1}^{S M}=\mathbf{x}_{k}^{S M}-t_{k}\left(\gamma_{k}^{S M}\right)^{-1} \mathbf{g}_{k} \tag{1.9}
\end{equation*}
$$

where $t_{k}>0$ is the basic step size and $\gamma_{k}^{S M}>0$ is the gain parameter determined as in

$$
\gamma_{k+1}^{S M}=2 \gamma_{k}^{S M} \frac{\gamma_{k}^{S M}\left[f\left(\mathbf{x}_{k+1}^{S M}\right)-f\left(\mathbf{x}_{k}^{S M}\right)\right]+t_{k}\left\|\mathbf{g}_{k}\right\|^{2}}{t_{k}^{2}\left\|\mathbf{g}_{k}\right\|^{2}}
$$

The ADSS model from [10] is defined as

$$
\begin{equation*}
\mathbf{x}_{k+1}^{A D S S}=\mathbf{x}_{k}^{A D S S}-\left(t_{k}\left(\gamma_{k}^{A D S S}\right)^{-1}+l_{k}\right) \mathbf{g}_{k} \tag{1.10}
\end{equation*}
$$

where $t_{k}$ and $l_{k}$ are determined by BLSs. The TADSS method [15] is defined by the iterative rule

$$
\mathbf{x}_{k+1}^{T A D S S}=\mathbf{x}_{k}^{T A D S S}-\left(t_{k}\left(\left(\gamma_{k}^{T A D S S}\right)^{-1}-1\right)+1\right) \mathbf{g}_{k}
$$

The next scheme was proposed as the modified SM (MSM) method in [7]:

$$
\begin{equation*}
\mathbf{x}_{k+1}^{M S M}=\mathbf{x}_{k}^{M S M}-\left(t_{k}+t_{k}^{2}-t_{k}^{3}\right)\left(\gamma_{k}^{M S M}\right)^{-1} \mathbf{g}_{k} \tag{1.11}
\end{equation*}
$$

The acceleration parameters in ADD, ADSS, TADSS and MSM methods are summarized in Table 1.1.

Table 1.1: Acceleration parameters $\gamma_{k+1}$ in variants SM method.

| Method | Acceleration parameter $\gamma_{k+1}$ | Reference |
| :---: | :---: | :---: |
| ADD | $\gamma_{k+1}^{A D D}=2 \frac{f\left(\mathbf{x}_{k+1}^{A D D}\right)-f\left(\mathbf{x}_{k}^{A D D}\right)-t_{k}\left(\mathbf{g}_{k}^{A D D}\right)^{1}\left(t_{k} \mathbf{d}_{k}^{A D D}-\left(\gamma_{k}\right)^{-1} \mathbf{g}_{k}\right)}{\left(t_{k} \mathbf{d}_{k}^{A D D}-\gamma_{k}^{-1} \mathbf{g}_{k}\right)^{1}\left(t_{k} \mathbf{d}_{k}^{A D D}-\left(\gamma_{k}^{A D D}\right)^{-1} \mathbf{g}_{k}\right)}$ | (2014) [11] |
| ADSS | $\gamma_{k+1}^{A D S S}=2 \frac{f\left(\mathbf{x}_{k+1}^{A D S}\right)-f\left(\mathbf{x}_{k}^{A} A S S\right)+\left(t_{k}\left(\gamma_{k}\right)^{-1}+l_{k}\right)\left\\|\mathbf{g}_{k}\right\\|^{2}}{\left(t_{k}\left(\gamma_{k}^{A D S S}\right)^{-1}+l_{k}\right)^{2}\left\\|\mathbf{g}_{k}\right\\|^{2}}$ | (2015) [10] |
| TADSS | $\begin{aligned} & \gamma_{k+1}^{T A D S S}=2 \frac{f\left(\mathbf{x}_{k+1}^{T A D S S}\right)-f\left(\mathbf{x}_{k}^{T A D S S}\right)+\psi_{k}\left\\|\mathbf{g}_{k}\right\\|^{2}}{\psi_{k}^{2}\left\\|\mathbf{g}_{k}\right\\|^{2}} \\ & \psi_{k}=t_{k}\left(\left(\gamma_{k}^{T A D S S}\right)^{-1}-1\right)+1 \end{aligned}$ | (2015) [15] |
| MSM | $\gamma_{k+1}^{M S M}=2 \gamma_{k} \frac{\left.\gamma_{k} \mid f\left(\mathbf{x}_{k+1}^{M S M}\right)-f\left(\mathbf{x}_{k}^{M S M}\right)\right]+\left(t_{k}+t_{k}^{2}-t_{k}^{3}\right)\left\\|\mathbf{g}_{k}\right\\|^{2}}{\left(t_{k}+t_{k}^{2}-t_{k}^{3}\right)^{2}\left\\|\mathbf{g}_{k}\right\\|^{2}}$ | (2019) [7] |

The main goal of this research is to study the impact of multiple usage of backtracking line search in modified SM method [7] and practical computational performance of two new methods. Our intention is to propose and investigate improvements of the MSM method. Globally, we investigate possibility to multiple use backtracking line search in the modified MSM method.

Main results of the present study can be highlighted as follows:
(1) A novel iterative scheme is proposed using the idea of computing the step parameters $t_{k}, t_{k}^{2}$ and $t_{k}^{3}$ in the MSM method by means of multiple BLS procedures. The resulting iterations will be denoted as TMSM and DMSM.
(2) Convergence behavior of the proposed iterations are investigated on appropriate quadratic functions.
(3) Numerical experiments compare introduced methods with existing iterations and analyze three main performances: number of iterative steps and function evaluations and CPU time.

The remainder of the paper is developed according to the following hierarchy of sections. Two modifications of the MSM methods, termed as TMSM and DMSM methods, are introduced in Section 2. Section 3. investigates the convergence of the presented TMSM and DMSM methods. In Section 4., we perform a number of numerical experiments and compare main performances of the novel methods with similar available methods. Final remarks are presented in Section 5.

## 2. Multiple use of backtracking line search in modified SM method

The MSM method is based on the iteration

$$
\begin{equation*}
\mathbf{x}_{k+1}^{M S M}=\mathbf{x}_{k}^{M S M}-t_{k}^{M S M}\left(\gamma_{k}^{M S M}\right)^{-1} \mathbf{g}_{k} \tag{2.1}
\end{equation*}
$$

where $t_{k}^{M S M}=t_{k}+t_{k}^{2}-t_{k}^{3}$. The leading idea in defining $t_{k}^{M S M}$ arises from the observation $t_{k}+t_{k}^{2}>t_{k}^{M S M}>t_{k}$, which means that the MSM method proposes a slightly greater step size with respect to the SM iterations. Since $t_{k}$ arises from the BLS procedure, which ensures $t_{k} \in(0,1)$, it implies

$$
t_{k} \leq t_{k}^{M S M} \leq t_{k}+t_{k}^{2}
$$

Our intention in current research is to improve behaviour of iterations (2.1) using two or three appropriately defined step-parameters. Following this idea, a method based on triple usage of the BLS in the MSM method is obtained when $t_{k}^{2}$ is substituted with $l_{k}^{2}$ and $t_{k}^{3}$ is substituted with $j_{k}^{3}$ in (2.1), where $t_{k}, l_{k}$ and $j_{k}$ are defined by independent LS procedures: the first BLS (Algorithm 1) calculates $t_{k}$, another BLS (Algorithm 2) calculates $l_{k}$, while the third BLS (Algorithm 3) determines $j_{k}$.

Replacing the above changes gives the expression of the TMSM iteration:

$$
\begin{equation*}
\mathbf{x}_{k+1}^{T M S M}=\mathbf{x}_{k}^{T M S M}-t_{k}^{T M S M}\left(\gamma_{k}^{T M S M}\right)^{-1} \mathbf{g}_{k} \tag{2.2}
\end{equation*}
$$

where

$$
t_{k}^{T M S M}= \begin{cases}t_{k}+l_{k}^{2}-j_{k}^{3}, & t_{k}+l_{k}^{2}-j_{k}^{3}>t_{k}  \tag{2.3}\\ t_{k}, & t_{k}+l_{k}^{2}-j_{k}^{3} \leq t_{k}\end{cases}
$$

The second order Taylor development of $f\left(\mathbf{x}_{k+1}^{T M S M}\right)$ gives

$$
\begin{align*}
f\left(\mathbf{x}_{k+1}^{T M S M}\right) \approx & f\left(\mathbf{x}_{k}^{T M S M}\right)-t_{k}^{T M S M}\left(\gamma_{k}^{T M S M}\right)^{-1} \mathbf{g}_{k}^{T} \mathbf{g}_{k} \\
& +\frac{1}{2}\left(t_{k}^{T M S M}\right)^{2}\left(\left(\gamma_{k}^{T M S M}\right)^{-1} \mathbf{g}_{k}\right)^{\mathrm{T}} \nabla^{2} f(\xi)\left(\gamma_{k}^{T M S M}\right)^{-1} \mathbf{g}_{k} \tag{2.4}
\end{align*}
$$

The parameter $\xi$ in (2.4) fulfills the condition $\xi \in\left[\mathbf{x}_{k}^{T M S M}, \mathbf{x}_{k+1}^{T M S M}\right]$. One possible choice is

$$
\begin{align*}
\xi & =\mathbf{x}_{k}^{T M S M}+\delta\left(\mathbf{x}_{k+1}^{T M S M}-\mathbf{x}_{k}^{T M S M}\right) \\
& =\mathbf{x}_{k}^{T M S M}-\varphi t_{k}^{T M S M}\left(\gamma_{k}^{T M S M}\right)^{-1} \mathbf{g}_{k}, \quad 0 \leq \varphi \leq 1 \tag{2.5}
\end{align*}
$$

According to [14], $\nabla^{2} f(\xi)$ is approximated as $\gamma_{k+1}^{T M S M} I$. So, (2.4) reduces to

$$
\begin{align*}
f\left(\mathbf{x}_{k+1}^{T M S M}\right) \approx & f\left(\mathbf{x}_{k}^{T M S M}\right)-t_{k}^{T M S M}\left(\gamma_{k}^{T M S M}\right)^{-1}\left\|\mathbf{g}_{k}\right\|^{2} \\
& +\frac{1}{2}\left(t_{k}^{T M S M}\right)^{2} \gamma_{k+1}^{T M S M}\left(\gamma_{k}^{T M S M}\right)^{-2}\left\|\mathbf{g}_{k}\right\|^{2} . \tag{2.6}
\end{align*}
$$

Then $\gamma_{k+1}^{T M S M}$ can be obtained from (2.6) as

$$
\begin{equation*}
\gamma_{k+1}^{T M S M}=2 \gamma_{k}^{T M S M} \frac{\gamma_{k}^{T M S M}\left[f\left(\mathbf{x}_{k+1}^{T M S M}\right)-f\left(\mathbf{x}_{k}^{T M S M}\right)\right]+t_{k}^{T M S M}\left\|\mathbf{g}_{k}\right\|^{2}}{\left(t_{k}^{T M S M}\right)^{2}\left\|\mathbf{g}_{k}\right\|^{2}} \tag{2.7}
\end{equation*}
$$

The improper situation $\gamma_{k+1}^{T M S M}<0$ can be resolved by taking $\gamma_{k+1}^{T M S M}=1$.
The BLS method is implemented in the Algorithm 1 from [14]. Algorithm 1 defines $t_{k}$ starting from $t=1$ and subsequently decreases values of $t$ so that it reduces the value of the objective $f$ enough.

```
Algorithm 1 The backtracking line search calculates \(t_{k}\).
Require: A real function \(f(\mathbf{x})\), appropriate search direction \(\mathbf{d}_{k}\) at the point \(\mathbf{x}_{k}\)
    and the positive real numbers \(0<\sigma<0.5\) and \(\beta \in(0,1)\).
    \(t=1\).
    While \(f\left(\mathbf{x}_{k}+t \mathbf{d}_{k}\right)>f\left(\mathbf{x}_{k}\right)+\sigma t \mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k}\), do \(t:=t \beta\).
    Output \(t_{k}:=t\).
```

```
Algorithm 2 The second backtracking line search calculates \(l_{k}\).
Require: Objective function \(f(\mathbf{x})\), the search direction \(\mathbf{d}_{k}\) at the point \(\mathbf{x}_{k}\) and
    positive real numbers \(0<\sigma_{l}<0.5\) and \(\beta_{l} \in(0,1)\).
    \(l=1\).
    While \(f\left(\mathbf{x}_{k}+l \mathbf{d}_{k}\right)>f\left(\mathbf{x}_{k}\right)+\sigma_{l} l \mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k}\), take \(l:=l \beta_{l}\).
    Return \(l_{k}=l\).
```

```
Algorithm 3 The third backtracking line search calculates \(j_{k}\).
Require: Objective function \(f(\mathbf{x})\), the search direction \(\mathbf{d}_{k}\) at the point \(\mathbf{x}_{k}\) and
    positive real numbers \(0<\sigma_{j}<0.5\) and \(\beta_{j} \in(0,1)\).
    \(j=1\).
    While \(f\left(\mathbf{x}_{k}+j \mathbf{d}_{k}\right)>f\left(\mathbf{x}_{k}\right)+\sigma_{j} j \mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k}\), take \(j:=j \beta_{j}\).
    Return \(j_{k}=j\).
```

Finally, the TMSM method is described in Algorithm 4.
It is expectable that the total number of iterations (NofI) required by the TMSM method will be smaller than the number of iterations of the MSM method, but an increase in the number of function evaluations (NofFE) and the CPU time (CPUT)is expectable. Based on these indicators, we came up with the idea to omit one line search in the TMSM method. This would drastically reduce the CPUT and the NofFE. Following this idea, a method of double use backtracking line search in modified SM method is obtained. In this way, we get a new expression of the DMSM iteration:

$$
\begin{equation*}
\mathbf{x}_{k+1}^{D M S M}=\mathbf{x}_{k}^{D M S M}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}, \tag{2.8}
\end{equation*}
$$

```
Algorithm 4 Triple use of backtracking line search in the MSM method (the TMSM method)
```

Require: Objective function $f(\mathbf{x})$, initial point $\mathbf{x}_{0}^{T M S M} \in \operatorname{dom}(f)$ and parameters $0<\lambda<1,0<\nu<1$.
1: Put $k=0$, evaluate $f\left(\mathbf{x}_{0}^{T M S M}\right), \mathbf{g}_{0}=\nabla f\left(\mathbf{x}_{0}^{T M S M}\right)$, and put $\gamma_{0}^{T M S M}=1$.
If

$$
\left\|\mathbf{g}_{k}\right\| \leq \lambda \quad \text { and } \quad \frac{\left|f\left(\mathbf{x}_{k+1}^{T M S M}\right)-f\left(\mathbf{x}_{k}^{T M S M}\right)\right|}{1+\left|f\left(\mathbf{x}_{k}^{T M S M}\right)\right|} \leq \nu
$$

STOP; else go to Step 3.
3: (The first backtracking) Compute $t_{k} \in(0,1]$ using Algorithm 1.
(The second backtracking) Compute $l_{k} \in(0,1]$ using Algorithm 2.
: (The third backtracking) Compute $j_{k} \in(0,1]$ using Algorithm 3.
: Determine $t_{k}^{T M S M}$ using (2.3).
Compute $\mathbf{x}_{k+1}^{T M S M}=\mathbf{x}_{k}^{T M S M}-\left(\gamma_{k}^{T M S M}\right)^{-1} t_{k}^{T M S M} \mathbf{g}_{k}$.
8: Compute $f\left(\mathbf{x}_{k+1}^{T M M}\right)$ and $\mathbf{g}_{k+1}=\nabla f\left(\mathbf{x}_{k+1}^{T M S M}\right)$.
9: Determine $\gamma_{k+1}^{T M S M}$ using (2.7).
10: If $\gamma_{k+1}^{T M S M}<0$, then take $\gamma_{k+1}^{T M S M}=1$.
11: Set $k:=k+1$, go to the step 2 .
12: Return $\mathbf{x}_{k+1}^{T M S M}$ and $f\left(\mathbf{x}_{k+1}^{T M S M}\right)$.
where

$$
t_{k}^{D M S M}= \begin{cases}t_{k}+t_{k}^{2}-j_{k}^{3}, & t_{k}+t_{k}^{2}-j_{k}^{3}>t_{k}  \tag{2.9}\\ t_{k}, & t_{k}+t_{k}^{2}-j_{k}^{3} \leq t_{k}\end{cases}
$$

In exactly the same way as for the TMSM method, we arrive at

$$
\begin{equation*}
\gamma_{k+1}^{D M S M}=2 \gamma_{k}^{D M S M} \frac{\gamma_{k}^{D M S M}\left[f\left(\mathbf{x}_{k+1}^{D M S M}\right)-f\left(\mathbf{x}_{k}^{D M S M}\right)\right]+t_{k}^{D M S M}\left\|\mathbf{g}_{k}\right\|^{2}}{\left(t_{k}^{T M S M}\right)^{2}\left\|\mathbf{g}_{k}\right\|^{2}} . \tag{2.10}
\end{equation*}
$$

The difficulty $\gamma_{k+1}^{D M S M}<0$ can be resolved using $\gamma_{k+1}^{D M S M}=1$.
The DMSM method is presented in Algorithm 5:

```
Algorithm 5 Double use backtracking line search in the MSM method (the DMSM
method)
```

Require: Function $f(\mathbf{x})$, chosen initial point $\mathbf{x}_{0}^{D M S M} \in \operatorname{dom}(f)$ and parameters $0<\lambda<1,0<\nu<1$.
1: Put $k=0$, evaluate $f\left(\mathbf{x}_{0}^{D M S M}\right), \mathbf{g}_{0}=\nabla f\left(\mathbf{x}_{0}^{D M S M}\right)$ and take $\gamma_{0}^{D M S M}=1$.
If

$$
\left\|\mathbf{g}_{k}\right\| \leq \lambda \quad \text { and } \quad \frac{\left|f\left(\mathbf{x}_{k+1}^{D M S M}\right)-f\left(\mathbf{x}_{k}^{D M S M}\right)\right|}{1+\left|f\left(\mathbf{x}_{k}^{D M S M}\right)\right|} \leq \nu
$$

STOP; else go to Step 3.
3: (The first backtracking) Compute $t_{k} \in(0,1]$ using Algorithm 1.
4: (The second backtracking) Compute $j_{k} \in(0,1]$ using Algorithm 3.
5: Determine $t_{k}^{D M S M}$ using (2.9).
6: Compute $\mathbf{x}_{k+1}^{D M S M}=\mathbf{x}_{k}^{D M S M}-\left(\gamma_{k}^{D M S M}\right)^{-1} t_{k}^{D M S M} \mathbf{g}_{k}$.
7: Compute $f\left(\mathbf{x}_{k+1}^{D M S M}\right)$ and $\mathbf{g}_{k+1}=\nabla f\left(\mathbf{x}_{k+1}^{D M S M}\right)$.
8: Determine the scalar approximation $\gamma_{k+1}^{D M S M} I$ of the Hessian of $f$ at the point $\mathbf{x}_{k+1}^{D M S M}$ using (2.10).
9: If $\gamma_{k+1}^{D M S M}<0$, then take $\gamma_{k+1}^{D M S M}=1$.
10: Put $k:=k+1$, go to the step 2.
11: Return $\mathbf{x}_{k+1}^{D M S M}$ and $f\left(\mathbf{x}_{k+1}^{D M S M}\right)$.

## 3. Convergence analysis

The content of this section is the convergence analysis of the TMSM and DMSM methods. In the following part, we restate and derive some basic statements which will be used in the convergence analysis of Algorithms 4 and 5 . The proofs can be found in $[1,9,12,13,14]$ and have been omitted:
$\left(H_{1}\right)$ the function $f$ is bounded below on $B_{0}=\left\{\mathbf{x} \in \mathbb{R}^{n} \mid f(\mathbf{x}) \leq f\left(\mathbf{x}_{0}\right)\right\}$;
$\left(H_{2}\right)$ the gradient $\mathbf{g}$ is Lipschitz continuous in an open convex set $B \supseteq B_{0}$ :

$$
\begin{equation*}
\|\mathbf{g}(\mathbf{x})-\mathbf{g}(\mathbf{y})\| \leq L\|\mathbf{x}-\mathbf{y}\|, \quad \forall \mathbf{x}, \mathbf{y} \in B, L>0 \tag{3.1}
\end{equation*}
$$

Proposition 3.1. $[1,13]$ Let $\mathbf{d}_{k}$ be a descent direction and the gradient $\mathbf{g}_{k}$ satisfies the Lipschitz condition (3.1). If $t_{k}$ is determined by the BLS in Algorithm 1, then

$$
\begin{equation*}
t_{k} \geq \min \left\{1,-\frac{\beta(1-\sigma)}{L} \frac{\mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k}}{\left\|\mathbf{d}_{k}\right\|^{2}}\right\} \tag{3.2}
\end{equation*}
$$

Lemma 3.1. If the function $f$ is $U C$ and $T C D$ on $\mathbb{R}^{n}$ then there exist $m, M$ such that

$$
\begin{equation*}
0<m \leq 1 \leq M \tag{3.3}
\end{equation*}
$$

then $f(\mathbf{x})$ possesses a minimizer $\mathbf{x}^{*}$ and

$$
\begin{array}{r}
m\|\mathbf{y}\|^{2} \leq \mathbf{y}^{\mathrm{T}} \nabla^{2} f(\mathbf{x}) \mathbf{y} \leq M\|\mathbf{y}\|^{2}, \quad \forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^{n} \\
\frac{1}{2} m\left\|\mathbf{x}-\mathbf{x}^{*}\right\|^{2} \leq f(\mathbf{x})-f\left(\mathbf{x}^{*}\right) \leq \frac{1}{2} M\left\|\mathbf{x}-\mathbf{x}^{*}\right\|^{2}, \quad \forall \mathbf{x} \in \mathbb{R}^{n} \tag{3.5}
\end{array}
$$

(3.6) $m\|\mathbf{x}-\mathbf{y}\|^{2} \leq(\mathbf{g}(\mathbf{x})-\mathbf{g}(\mathbf{y}))^{\mathrm{T}}(\mathbf{x}-\mathbf{y}) \leq M\|\mathbf{x}-\mathbf{y}\|^{2}, \quad \forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^{n}$.

Lemma 3.2. [14] The following inequality holds for a TCD and UC function $f$ and for the IGD sequence $\left\{\mathbf{x}_{k}\right\}$ generated by (1.6):

$$
\begin{equation*}
f\left(\mathbf{x}_{k}\right)-f\left(\mathbf{x}_{k+1}\right) \geq \mu\left\|\mathbf{g}_{k}\right\|^{2} \tag{3.7}
\end{equation*}
$$

with

$$
\begin{equation*}
\mu=\min \left\{\frac{\sigma}{M}, \frac{\sigma(1-\sigma)}{L} \beta\right\} . \tag{3.8}
\end{equation*}
$$

In further, it is assumed in this section that $\mathbf{d}_{k}$ is a descent direction. Further, the scalar approximation of Hessian is TCD. Moreover, instead of (3.4) and (3.3) it is sufficient to assume:

$$
\begin{equation*}
m \leq \gamma_{k} \leq M, \quad 0<m \leq 1 \leq M, \quad m, M \in \mathbb{R} \tag{3.9}
\end{equation*}
$$

So, all values $\gamma_{k}<0$ will be replaced by $\gamma_{k}=1$, while the cases $\gamma_{k}>M$ will be resolved by $\gamma_{k}=M$.

Theorem 3.1. Let $\left(H_{1}\right)$ and $\left(H_{2}\right)$ and (3.9) be true and the mapping $f$ is UC. Then the sequence $\left\{\mathbf{x}_{k}^{D M S M}\right\}$ fulfils (3.7)-(3.8).

Proof. From (2.8), it can be concluded

$$
\begin{aligned}
\mathbf{x}_{k+1}^{D M S M} & =\mathbf{x}_{k}^{D M S M}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k} \\
& =\mathbf{x}_{k}^{D M S M}-t_{k} \frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k} \\
& =\mathbf{x}_{k}^{D M S M}+t_{k} \mathbf{d}_{k}
\end{aligned}
$$

where $\mathbf{d}_{k}=-\frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}$.
Based on the stopping condition of the backtracking line search (Algorithm 1), we conclude

$$
\begin{equation*}
f\left(\mathbf{x}_{k}^{D M S M}\right)-f\left(\mathbf{x}_{k+1}^{D M S M}\right) \geq-\sigma t_{k} \mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k} . \quad \forall k \in \mathbb{N} . \tag{3.10}
\end{equation*}
$$

In the situation $t_{k}<1$, by putting expression for $\mathbf{d}_{k}$ into (3.10), the following inequalities can be derived:

$$
\begin{align*}
f\left(\mathbf{x}_{k}^{D M S M}\right)-f\left(\mathbf{x}_{k+1}^{D M S M}\right) & \geq-\sigma t_{k} \mathbf{g}_{k}{ }^{\mathrm{T}} \mathbf{d}_{k} \\
& =-\sigma t_{k} \mathbf{g}_{k}^{\mathrm{T}}\left(-\frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right)  \tag{3.11}\\
& =\sigma t_{k} \frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1}\left\|\mathbf{g}_{k}\right\|^{2} .
\end{align*}
$$

Now, from (3.2), it follows that

$$
\begin{align*}
& t_{k} \geq-\frac{\beta(1-\sigma)}{L} \cdot \frac{\mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k}}{\left\|\mathbf{d}_{k}\right\|^{2}} \\
&\left.=-\frac{\beta(1-\sigma)}{L} \cdot \frac{\mathbf{g}_{k}^{\mathrm{T}}\left(-\frac{t_{k}^{D M S M}}{t_{k}}\right.}{}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right) \\
&=\frac{\beta(1-\sigma)}{L} \cdot \frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k} \|^{2}  \tag{3.12}\\
&\left.=\frac{\beta(1-\sigma)}{L} \cdot \frac{\mathbf{t}_{k}^{D M S M}}{t_{k}}\right)^{2}\left(\gamma_{k}^{D M S M}\right. \\
&\left.\frac{t_{k}}{D M S M}\right)^{-2}\left\|\gamma_{k}^{D M S M}\right\|^{-1} \mathbf{g}_{k} \\
&=\frac{(1-\sigma) \beta}{L} \cdot \frac{t_{k} \|_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1}\left\|\mathbf{g}_{k}\right\|^{2} \\
& t_{k}^{D M S M}
\end{align*}
$$

By applying inequality (3.12) to (3.11), we obtain

$$
\begin{align*}
& f\left(\mathbf{x}_{k}^{D M S M}\right)-f\left(\mathbf{x}_{k+1}^{D M S M}\right) \geq \sigma t_{k} \frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1}\left\|\mathbf{g}_{k}\right\|^{2} \\
& \geq \sigma \frac{(1-\sigma) \beta}{L} \cdot \frac{\gamma_{k}^{D M S M}}{\frac{t_{k}^{D M S M}}{t_{k}}} \frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1}\left\|\mathbf{g}_{k}\right\|^{2}  \tag{3.13}\\
& \geq \sigma \frac{(1-\sigma) \beta}{L}\left\|\mathbf{g}_{k}\right\|^{2} .
\end{align*}
$$

In the case $t_{k}=1$, based on (3.9) and (3.10) the following inequality holds

$$
\begin{align*}
f\left(\mathbf{x}_{k}^{D M S M}\right)-f\left(\mathbf{x}_{k+1}^{D M S M}\right) & \geq-\sigma \mathbf{g}_{k}^{\mathrm{T}} \mathbf{d}_{k} \\
& =-\sigma \mathbf{g}_{k}^{\mathrm{T}}\left(-\frac{t_{k}^{D M S M}}{t_{k}}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right)  \tag{3.14}\\
& =\frac{\sigma}{\gamma_{k}^{D M S M}} \frac{t_{k}^{D M S M}}{t_{k}}\left\|\mathbf{g}_{k}\right\|^{2} .
\end{align*}
$$

According to (2.9), it follows that $t_{k}^{D M S M} \geq t_{k}$, which implies

$$
\begin{align*}
f\left(\mathbf{x}_{k}^{D M S M}\right)-f\left(\mathbf{x}_{k+1}^{D M S M}\right) & \geq \frac{\sigma}{\gamma_{k}^{D M S M}}\left\|\mathbf{g}_{k}\right\|^{2}  \tag{3.15}\\
& \geq \frac{\sigma}{M}\left\|\mathbf{g}_{k}\right\|^{2} .
\end{align*}
$$

Finally, from (3.13) and (3.15) we get (3.8).

Theorem 3.2. Let $\left(H_{1}\right)$ and $\left(H_{2}\right)$ are valid in conjunction with (3.9) and $f$ be a UC function.
(a) The sequence $\left\{\mathbf{x}_{k}^{D M S M}\right\}$ satisfies $\lim _{k \rightarrow \infty}\left\|\mathbf{g}_{k}\right\|=0$, and $\left\{\mathbf{x}_{k}^{D M S M}\right\}$ converges to $\mathbf{x}^{*}$.
(b) The sequence $\left\{\mathbf{x}_{k}^{T M S M}\right\}$ satisfies $\lim _{k \rightarrow \infty}\left\|\mathbf{g}_{k}\right\|=0$, and $\left\{\mathbf{x}_{k}^{T M S M}\right\}$ converges to $\mathbf{x}^{*}$.

Proof. Analogously as the proof of [14, Theorem 4.1].

Lemma 3.3 confirms the convergence of the DMSM method on the strictly convex quadratic (SCQ) functions

$$
\begin{equation*}
f(\mathbf{x})=\frac{1}{2} \mathbf{x}^{\mathrm{T}} A \mathbf{x}-\mathbf{b}^{\mathrm{T}} \mathbf{x} \tag{3.16}
\end{equation*}
$$

where $A \in \mathbb{R}^{n \times n}$ is a symmetric positive definite matrix and $\mathbf{b} \in \mathbb{R}^{n}$. The eigenvalues of $A$ are ordered as $\lambda_{1} \leq \cdots \leq \lambda_{n}$.

Lemma 3.3. The DMSM iterations (2.8) applied on a SCQ function $f$ given by the expression (3.16) satisfy the inequality

$$
\begin{equation*}
\lambda_{1} \leq \frac{\gamma_{k+1}^{D M S M}}{t_{k+1}} \leq \frac{2 \lambda_{n}}{\sigma}, k \in \mathbb{N} \tag{3.17}
\end{equation*}
$$

Proof. Simple verification gives

$$
\begin{align*}
f\left(\mathbf{x}_{k+1}^{D M S M}\right)-f\left(\mathbf{x}_{k}^{D M S M}\right)= & \frac{1}{2}\left(\mathbf{x}_{k+1}^{D M S M}\right)^{\mathrm{T}} A \mathbf{x}_{k+1}^{D M S M}-\mathbf{b}^{\mathrm{T}} \mathbf{x}_{k+1}^{D M S M} \\
& -\frac{1}{2}\left(\mathbf{x}_{k}^{D M S M}\right)^{\mathrm{T}} A \mathbf{x}_{k}^{D M S M}+\mathbf{b}^{\mathrm{T}} \mathbf{x}_{k}^{D M S M} . \tag{3.18}
\end{align*}
$$

The substitute of (2.8) in (3.18) gives

$$
\begin{align*}
f\left(\mathbf{x}_{k+1}^{D M S M}\right)- & f\left(\mathbf{x}_{k}^{D M S M}\right)=\frac{1}{2}\left[\mathbf{x}_{k}^{D M S M}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right]^{\mathrm{T}} \\
& \times A\left[\mathbf{x}_{k}^{D M S M}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right] \\
& -\mathbf{b}^{\mathrm{T}}\left[\mathbf{x}_{k}^{D M S M}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right] \\
& -\frac{1}{2}\left(\mathbf{x}_{k}^{D M S M}\right)^{\mathrm{T}} A \mathbf{x}_{k}^{D M S M}+\mathbf{b}^{\mathrm{T}} \mathbf{x}_{k}^{D M S M} \\
= & -\frac{1}{2} t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1}\left(\mathbf{x}_{k}^{D M S M}\right)^{\mathrm{T}} A \mathbf{g}_{k}  \tag{3.19}\\
& -\frac{1}{2} t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}^{\mathrm{T}} A \mathbf{x}_{k}^{D M S M} \\
& +\frac{1}{2}\left(t_{k}^{D M S M}\right)^{2}\left(\gamma_{k}^{D M S M}\right)^{-2} \mathbf{g}_{k}^{\mathrm{T}} A \mathbf{g}_{k} \\
& +t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{b}^{\mathrm{T}} \mathbf{g}_{k} .
\end{align*}
$$

Since the gradient of the function (3.16) corresponding to DMSM is equal to

$$
\begin{equation*}
\mathbf{g}_{k}=A \mathbf{x}_{k}^{D M S M}-\mathbf{b} \tag{3.20}
\end{equation*}
$$

one can verify

$$
\begin{align*}
f\left(\mathbf{x}_{k+1}^{D M S M}\right)- & f\left(\mathbf{x}_{k}^{D M S M}\right) \\
= & t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1}\left[\mathbf{b}^{\mathrm{T}} \mathbf{g}_{k}-\left(\mathbf{x}_{k}^{D M S M}\right)^{\mathrm{T}} A \mathbf{g}_{k}\right] \\
& +\frac{1}{2}\left(t_{k}^{D M S M}\right)^{2}\left(\gamma_{k}^{D M S M}\right)^{-2} \mathbf{g}_{k}^{\mathrm{T}} A \mathbf{g}_{k} \\
= & t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1}\left[\mathbf{b}^{\mathrm{T}}-\left(\mathbf{x}_{k}^{D M S M}\right)^{\mathrm{T}} A\right] \mathbf{g}_{k}  \tag{3.21}\\
& +\frac{1}{2}\left(t_{k}^{D M S M}\right)^{2}\left(\gamma_{k}^{D M S M}\right)^{-2} \mathbf{g}_{k}^{\mathrm{T}} A \mathbf{g}_{k} \\
= & -t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}^{\mathrm{T}} \mathbf{g}_{k} \\
& +\frac{1}{2}\left(t_{k}^{D M S M}\right)^{2}\left(\gamma_{k}^{D M S M}\right)^{-2} \mathbf{g}_{k}^{\mathrm{T}} A \mathbf{g}_{k} .
\end{align*}
$$

After substitute (3.21) into (2.10), the parameter $\gamma_{k+1}^{D M S M}$ becomes

$$
\begin{align*}
\gamma_{k+1}^{D M S M} & =2 \gamma_{k}^{D M S M} \frac{\gamma_{k}^{D M S M}\left[f\left(\mathbf{x}_{k+1}^{D M S M}\right)-f\left(\mathbf{x}_{k}^{D M S M}\right)\right]+t_{k}^{D M S M}\left\|\mathbf{g}_{k}\right\|^{2}}{\left(t_{k}^{D M S M}\right)^{2}\left\|\mathbf{g}_{k}\right\|^{2}}  \tag{3.22}\\
& =\frac{\mathbf{g}_{k}^{\mathrm{T}} A \mathbf{g}_{k}}{\left\|\mathbf{g}_{k}\right\|^{2}} .
\end{align*}
$$

Therefore, the following inequalities are valid:

$$
\begin{equation*}
\lambda_{1} \leq \gamma_{k+1}^{D M S M} \leq \lambda_{n}, k \in \mathbb{N} \tag{3.23}
\end{equation*}
$$

The inequality in (3.17) follows from (3.23) in conjunction with $0<t_{k+1} \leq 1$. In order to verify the right hand side inequality in (3.17), it suffices to observe the upper bound caused by the BLS

$$
t_{k} \geq \frac{\beta(1-\sigma) \gamma_{k}}{L}
$$

which implies

$$
\begin{equation*}
\frac{\gamma_{k+1}^{D M S M}}{t_{k+1}}<\frac{L}{\beta(1-\sigma)} \tag{3.24}
\end{equation*}
$$

Using $\mathbf{g}(\mathbf{x})=A \mathbf{x}-\mathbf{b}$ in common with the fact that $A$ symmetric, it follows that

$$
\begin{equation*}
\|\mathbf{g}(\mathbf{x})-\mathbf{g}(\mathbf{y})\|=\|A \mathbf{x}-A \mathbf{y}\|=\|A(\mathbf{x}-\mathbf{y})\| \leq\|A\|\|\mathbf{x}-\mathbf{y}\|=\lambda_{n}\|\mathbf{x}-\mathbf{y}\| \tag{3.25}
\end{equation*}
$$

The Lipschitz constant $L$ in (3.24) can be equal to the largest eigenvalue $\lambda_{n}$. Using $\sigma \in(0,0.5), \beta \in(\sigma, 1)$ one obtains

$$
\begin{equation*}
\frac{\gamma_{k+1}^{D M S M}}{t_{k+1}}<\frac{L}{\beta(1-\sigma)}=\frac{\lambda_{n}}{\beta(1-\sigma)}<\frac{2 \lambda_{n}}{\sigma} . \tag{3.26}
\end{equation*}
$$

So, the right inequality in (3.17) is verified.

Theorem 3.3. Let $f$ be a $S C Q$ function defined in (3.16). In the case $\lambda_{n}<2 \lambda_{1}$ the DMSM method (2.8) satisfies

$$
\begin{equation*}
\left(d_{i}^{k+1}\right)^{2} \leq \delta^{2}\left(d_{i}^{k}\right)^{2} \tag{3.27}
\end{equation*}
$$

where

$$
\begin{equation*}
\delta=\max \left\{1-\frac{\sigma \lambda_{1}}{2 \lambda_{n}}, \frac{\lambda_{n}}{\lambda_{1}}-1\right\} \tag{3.28}
\end{equation*}
$$

In addition,

$$
\begin{equation*}
\lim _{k \rightarrow \infty}\left\|\mathbf{g}_{k}\right\|=0 \tag{3.29}
\end{equation*}
$$

Proof. Let $\left\{v_{1}, \ldots, v_{n}\right\}$ be orthonormal eigenvectors of $A$. On the basis of (3.20), there exist real quantities $d_{1}^{k}, d_{2}^{k}, \ldots, d_{n}^{k}$ satisfying

$$
\begin{equation*}
\mathbf{g}_{k}=\sum_{i=1}^{n} d_{i}^{k} v_{i} \tag{3.30}
\end{equation*}
$$

Now, using (2.8) one can simply deduce

$$
\begin{aligned}
\mathbf{g}_{k+1} & =A \mathbf{x}_{k+1}^{D M S M}-\mathbf{b} \\
& =A\left(\mathbf{x}_{k}^{D M S M}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \mathbf{g}_{k}\right)-\mathbf{b} \\
& =\mathbf{g}_{k}-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} A \mathbf{g}_{k} \\
& =\left(I-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} A\right) \mathbf{g}_{k} .
\end{aligned}
$$

Using the simple linear approximation of $\mathbf{g}_{k+1}$ as in (3.30), we get

$$
\begin{equation*}
\mathbf{g}_{k+1}=\sum_{i=1}^{n} d_{i}^{k+1} v_{i}=\sum_{i=1}^{n}\left(1-t_{k}^{D M S M}\left(\gamma_{k}^{D M S M}\right)^{-1} \lambda_{i}\right) d_{i}^{k} v_{i} \tag{3.31}
\end{equation*}
$$

To prove (3.27), it is enough to show that $\left|1-\frac{\lambda_{i}}{\gamma_{k}^{D M S M}\left(t_{k}^{D M S M}\right)^{-1}}\right| \leq \delta$. Two cases can be observed. First, if $\lambda_{i} \leq \frac{\gamma_{k}^{D M S M}}{t_{k}^{D M S M}}$ implying (3.17), we can conclude the following:

$$
\begin{equation*}
1>\frac{\lambda_{i}}{\gamma_{k}^{D M S M}\left(t_{k}^{D M S M}\right)^{-1}} \geq \frac{\sigma \lambda_{1}}{2 \lambda_{n}} \Longrightarrow 1-\frac{\lambda_{i}}{\gamma_{k}^{D M S M}\left(t_{k}^{D M S M}\right)^{-1}} \leq 1-\frac{\sigma \lambda_{1}}{2 \lambda_{n}} \leq \delta \tag{3.32}
\end{equation*}
$$

Now, let us examine another case $\frac{\gamma_{k}^{D M S M}}{t_{k}^{D M S M}}<\lambda_{i}$. Since

$$
\begin{equation*}
1<\frac{\lambda_{i}}{\gamma_{k}^{D M S M}\left(t_{k}^{D M S M}\right)^{-1}} \leq \frac{\lambda_{n}}{\lambda_{1}}, \tag{3.33}
\end{equation*}
$$

it follows that

$$
\begin{equation*}
\left|1-\frac{\lambda_{i}}{\gamma_{k}^{D M S M}\left(t_{k}^{D M S M}\right)^{-1}}\right| \leq \frac{\lambda_{n}}{\lambda_{1}}-1 \leq \delta \tag{3.34}
\end{equation*}
$$

Now, in order to prove $\lim _{k \rightarrow \infty}\left\|\mathbf{g}_{k}\right\|=0$, it suffices to use the orthonormality of $\left\{v_{1}, \ldots, v_{n}\right\}$ in common with (3.30) and conclude

$$
\begin{equation*}
\left\|\mathbf{g}_{k}\right\|^{2}=\sum_{i=1}^{n}\left(d_{i}^{k}\right)^{2} \tag{3.35}
\end{equation*}
$$

Since (3.27) is valid and $0<\delta<1$ holds, (3.35) initiates that (3.30).

## 4. Numerical results

All the considered methods are coded in Matlab R2017a programming language and executed on the notebook with Intel Core i3 2.0 GHz CPU, 8 GB RAM and Windows 10 operating system. The number of iterations (NofI), number of function evaluations (NofFE) and the CPU time (CPUT) are analyzed in numerical experiments.

Numerical testing is based on 24 test functions from [2], where a lot of the problems are taken over from CUTEr collection [4]. For each of tested functions in Tables 4.1, 4.2 and 4.3, 12 numerical testings are performed with $100,200,300,500$, $1000,2000,3000,5000,7000,8000,10000$ and 15000 unknowns. Tables 4.1, 4.2 and 4.3 arrange summary numerical results for AGD, MAGD, MSM, SM, DMSM and TMSM, tested on 24 functions.

For each of six tested methods (AGD, MAGD, SM, MSM, DMSM and TMSM), the same stopping criteria are used:

$$
\left\|\mathbf{g}_{k}\right\| \leq 10^{-6} \quad \text { and } \quad \frac{\left|f\left(\mathbf{x}_{k+1}\right)-f\left(\mathbf{x}_{k}\right)\right|}{1+\left|f\left(\mathbf{x}_{k}\right)\right|} \leq 10^{-16}
$$

The BLS parameters for AGD, MAGD, MSM and SM methods are $\sigma=0.0001$ and $\beta=0.8$. The backtracking procedures in the DMSM method are implemented using $\sigma=0.0001$ and $\beta=0.8$ for Algorithm 1 and $\sigma_{j}=0.00015$ and $\beta_{j}=0.85$ for Algorithm 3.

The backtracking procedures in the TMSM method are developed using $\sigma=$ 0.0001 and $\beta=0.8$ for Algorithm 1, $\sigma_{l}=0.0002$ and $\beta_{l}=0.9$ for Algorithm 2 and $\sigma_{j}=0.00015$ and $\beta_{j}=0.85$ for Algorithm 3.

Table 4.4 contains average values of NofI, the NofFE and the CPUT for all 288 numerical experiments.

Based on the values for NofI given in Table 4.4, it can be concluded that the DMSM and TMSM methods gives superior results with respect to MAGD, AGD, MSM and SM methods.

Table 4.1: Numerical results of the AGD, MAGD, MSM, SM, DMSM and TMSM methods for the NofI.

| Test function | MAGD | TMSM | MSM | DMSM | SM | AGD |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Perturbed Quadratic | 352325 | 31269 | 34828 | 31386 | 59908 | 353897 |
| Raydan 1 | 58504 | 30148 | 26046 | 17238 | 14918 | 22620 |
| Diagonal 3 | 119719 | 6767 | 7030 | 7077 | 12827 | 120416 |
| Generalized Tridiagonal 1 | 647 | 332 | 346 | 350 | 325 | 670 |
| Extended Tridiagonal 1 | 692219 | 685 | 1370 | 728 | 4206 | 3564 |
| Extended TET | 455 | 191 | 156 | 156 | 156 | 443 |
| Diagonal 4 | 8084 | 96 | 96 | 96 | 96 | 120 |
| Diagonal 5 | 48 | 72 | 72 | 72 | 72 | 48 |
| Extended Himmelblau | 302 | 312 | 260 | 264 | 196 | 396 |
| Perturbed quadratic diagonal | 1060824 | 36640 | 37454 | 31662 | 44903 | 2542050 |
| Quadratic QF1 | 362896 | 32099 | 36169 | 33138 | 62927 | 366183 |
| Extended quadratic penalty QP1 | 229 | 338 | 369 | 298 | 271 | 210 |
| Extended quadratic penalty QP2 | 356357 | 1735 | 1674 | 990 | 3489 | 395887 |
| Quadratic QF2 | 71647 | 31745 | 32727 | 30642 | 64076 | 100286 |
| Extended Tridiagonal 2 | 1665 | 694 | 659 | 583 | 543 | 1657 |
| ARWHEAD (CUTE) | 12834 | 328 | 430 | 302 | 270 | 5667 |
| Almost Perturbed Quadratic | 354369 | 30790 | 33652 | 32902 | 60789 | 356094 |
| LIARWHD (CUTE) | 925138 | 1257 | 3029 | 1726 | 18691 | 1054019 |
| ENGVAL1 (CUTE) | 822 | 623 | 461 | 434 | 375 | 743 |
| QUARTC (CUTE) | 177 | 302 | 217 | 220 | 290 | 171 |
| Generalized Quartic | 229 | 191 | 181 | 186 | 189 | 187 |
| Diagonal 7 | 159 | 144 | 147 | 111 | 108 | 72 |
| Diagonal 8 | 154 | 120 | 120 | 109 | 118 | 60 |
| Full Hessian FH3 | 63 | 63 | 63 | 63 | 63 | 45 |

Performance profiles from [6] are used in comparing the selected methods. As usual, the NofI, NofFE and CPUT profiles are used. All numerical results are represented in Figures 4.1 and 4.2. Figure 4.1 (left) shows the performances of compared methods related to NofI. Figure 4.1 (right) illustrates the performance of these methods relative to NofFE. Graphs in Figure 4.2 illustrate the behavior of considered methods with respect foto CPUT from the results displayet in fables 4.2 and 4.3 and according to graphs in Figures 4.1 and Figure 4.2, the following can be observed.
(1) The DMSM and TMSM methods give better results compared to other methods when we compare the number of iterations.
(2) The SM, MSM, DMSM and TMSM exhibit better performances than the AGD and MAGD methods.

From Figure 4.1 (left), it is observable that the graph of the DMSM method comes first to the top, which signifies that the DMSM outperforms other considered methods with respect to the NofI.

Table 4.2: Numerical results of the AGD, MAGD, MSM, SM, DMSM and TMSM methods for the NofFE.

| Test function | MAGD | TMSM | MSM | DMSM | SM | AGD |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Perturbed Quadratic | 13855459 | 645704 | 200106 | 370595 | 337910 | 13916515 |
| Raydan 1 | 1282162 | 1305952 | 311260 | 326766 | 81412 | 431804 |
| Diagonal 3 | 4244404 | 131307 | 38158 | 80193 | 69906 | 4264718 |
| Generalized Tridiagonal 1 | 9057 | 2934 | 1191 | 2061 | 1094 | 9334 |
| Extended Tridiagonal 1 | 2077341 | 14797 | 10989 | 9147 | 35621 | 14292 |
| Extended TET | 4130 | 1689 | 528 | 948 | 528 | 3794 |
| Diagonal 4 | 133440 | 2316 | 636 | 1320 | 636 | 1332 |
| Diagonal 5 | 108 | 300 | 156 | 228 | 156 | 108 |
| Extended Himmelblau | 5192 | 3636 | 976 | 1908 | 668 | 6897 |
| Perturbed quadratic diagonal | 38728371 | 1309740 | 341299 | 629088 | 460028 | 94921578 |
| Quadratic QF1 | 13192789 | 661661 | 208286 | 392426 | 352975 | 13310016 |
| Extended quadratic penalty QP1 | 2939 | 6400 | 2196 | 5421 | 2326 | 2613 |
| Extended quadratic penalty QP2 | 8846145 | 44962 | 11491 | 14058 | 25905 | 9852040 |
| Quadratic QF2 | 2810965 | 642829 | 183142 | 364257 | 353935 | 3989239 |
| Extended Tridiagonal 2 | 9613 | 9779 | 2866 | 4951 | 2728 | 8166 |
| ARWHEAD (CUTE) | 468970 | 15416 | 5322 | 8503 | 3919 | 214284 |
| Almost Perturbed Quadratic | 13936462 | 639129 | 194876 | 393591 | 338797 | 14003318 |
| LIARWHD (CUTE) | 41619197 | 39788 | 27974 | 33271 | 180457 | 47476667 |
| ENGVAL1 (CUTE) | 8332 | 10120 | 2285 | 4319 | 2702 | 6882 |
| QUARTC (CUTE) | 414 | 1412 | 494 | 780 | 640 | 402 |
| Generalized Quartic | 1244 | 1311 | 493 | 836 | 507 | 849 |
| Diagonal 7 | 745 | 930 | 504 | 696 | 335 | 333 |
| Diagonal 8 | 740 | 805 | 383 | 546 | 711 | 304 |
| Full Hessian FH3 | 1955 | 2160 | 566 | 1263 | 631 | 1352 |

Figure 4.1 (right) confirms that all six methods are able to solve all test cases. Further, the MSM method is superior in $58.33 \%$ of all tests with respect to MAGD ( $4.17 \%$ ), $\operatorname{TMSM}(0 \%), \operatorname{DMSM}(4.17 \%), \operatorname{SM}(29.17 \%)$ and $\operatorname{AGD}(16.67 \%)$.

Graphs in Figure 4.2 again confirm that all the methods are able to solve test problems, and the MSM is winer in $54.17 \%$ of the tests with respect to MAGD (4.17\%), $\operatorname{TMSM}(0 \%), \operatorname{DMSM}(4.17 \%), \operatorname{SM}(37.50 \%)$ and $\operatorname{AGD}(4.17 \%)$.

According to individual data arranged in the tables 4.1-4.3, generated average values as well as the presented graphs, the conclusion is that the DMSM method is winer concerning the NofI.

Compared to the previous numerical results obtained during the testing of AGD, MAGD, MSM, SM, DMSM and TMSM methods, in the next test for parameter values in the second and third backtracking line search we take the values that are less than the values in primary backtracking. The aim of this test is to answer the question: Does the choice of higher or lower parameter values in the second and third backtracking line search in relation to the primary backtracking line search

Table 4.3: Numerical results of the AGD, MAGD, MSM, SM, DMSM and TMSM methods for the CPUT.

| Test function | $M A G D$ | $T M S M$ | $M S M$ | $D M S M$ | $S M$ | $A G D$ |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Perturbed Quadratic | 6049.531 | 344.172 | 116.281 | 198.328 | 185.641 | 6756.047 |
| Raydan 1 | 334.266 | 388.156 | 31.906 | 67.344 | 36.078 | 158.359 |
| Diagonal 3 | 6401.969 | 199.547 | 52.609 | 120.406 | 102.875 | 5527.844 |
| Generalized Tridiagonal 1 | 7.781 | 4.641 | 1.469 | 3.625 | 1.203 | 11.344 |
| Extended Tridiagonal 1 | 8853.172 | 26.828 | 29.047 | 17.297 | 90.281 | 55.891 |
| Extended TET | 2.766 | 1.703 | 0.516 | 1.203 | 0.594 | 3.219 |
| Diagonal 4 | 16.172 | 0.719 | 0.203 | 0.359 | 0.141 | 0.781 |
| Diagonal 5 | 0.313 | 0.750 | 0.344 | 0.734 | 0.328 | 0.391 |
| Extended Himmelblau | 1.031 | 1.094 | 0.297 | 0.703 | 0.188 | 1.953 |
| Perturbed quadratic diagonal | 22820.172 | 534.750 | 139.625 | 273.188 | 185.266 | 44978.750 |
| Quadratic QF1 | 6846.453 | 258.938 | 81.531 | 168.453 | 138.172 | 12602.563 |
| Extended quadratic penalty QP1 | 1.063 | 2.234 | 1.000 | 3.516 | 0.797 | 1.266 |
| Extended quadratic penalty QP2 | 1872.797 | 12.578 | 3.516 | 8.063 | 6.547 | 3558.734 |
| Quadratic QF2 | 768.563 | 243.938 | 73.438 | 153.109 | 132.703 | 1582.766 |
| Extended Tridiagonal 2 | 2.531 | 4.938 | 1.047 | 2.375 | 1.031 | 3.719 |
| ARWHEAD (CUTE) | 138.000 | 6.422 | 1.969 | 4.609 | 1.359 | 95.641 |
| Almost Perturbed Quadratic | 7086.563 | 285.563 | 73.047 | 153.891 | 133.516 | 13337.125 |
| LIARWHD (CUTE) | 15372.625 | 10.203 | 9.250 | 12.641 | 82.016 | 27221.516 |
| ENGVAL1 (CUTE) | 2.641 | 4.328 | 1.047 | 2.375 | 1.188 | 3.906 |
| QUARTC (CUTE) | 2.078 | 4.531 | 1.844 | 3.297 | 2.313 | 2.469 |
| Generalized Quartic | 0.500 | 0.734 | 0.281 | 0.375 | 0.188 | 0.797 |
| Diagonal 7 | 0.688 | 0.953 | 0.547 | 1.469 | 0.375 | 0.625 |
| Diagonal 8 | 0.656 | 0.781 | 0.469 | 1.078 | 0.797 | 0.438 |
| Full Hessian FH3 | 1.188 | 1.672 | 0.391 | 1.234 | 0.391 | 1.438 |

Table 4.4: Average numerical outcomes for 24 test functions tested on 12 numerical experiments.

| Average performances | MAGD | TMSM | MSM | DMSM | SM | AGD |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| Number of iterations | 182494.42 | 8622.54 | 9064.83 | 7947.21 | 14575.25 | 221896.04 |
| No. of fun.evaluation | 5885007.25 | 228961.54 | 64424.04 | 110298.83 | 93938.63 | 8434868.21 |
| CPU time (sec) | 3190.98 | 97.51 | 25.90 | 49.99 | 46.00 | 4829.4 |

directly affect the numerical results of DMSM and TMSM methods?
The primary BLS uses the same parameters $\sigma=0.0001$ and $\beta=0.8$ as in the first test for AGD, MAGD, MSM and SM methods. The BLS procedures in the DMSM method are implemented using $\sigma=0.0001$ and $\beta=0.8$ for Algorithm 1 and $\sigma_{j}=0.00005$ and $\beta_{j}=0.7$ for Algorithm 3. Also, the BLS in the TMSM method are implemented using $\sigma=0.0001$ and $\beta=0.8$ for Algorithm $1, \sigma_{l}=0.00001$ and $\beta_{l}=0.6$


Fig. 4.1: Performance profiles based on the NofI (left) and NofFE (right).


Fig. 4.2: Performance profiles based upon CPUT.
for Algorithm 2 and $\sigma_{j}=0.00005$ and $\beta_{j}=0.7$ for Algorithm 3.
All other conditions (stop criteria and number of variables) remain the same as in the first numerical experiment.

The obtained numerical results are shown in the Tables 4.5, 4.6 and 4.7.
Table 4.8 includes the average values of NofI, the NofFE and the CPUT in a second numerical experiment.

According to the NofI values given in Table 4.8, it can be notified that the DMSM method gives better results and in the second numerical experiment compared to MAGD, AGD, MSM, SM and TMSM methods.

All numerical results from Tables 4.5, 4.6 and 4.7 are represented in Figures 4.3 and 4.4. Figure 4.3 (left) shows the NofI performances of compared methods. Figure 4.3 (right) demonstrates the NofFE profile of these methods. Figure 4.4

Table 4.5: Numerical results of the AGD, MAGD, MSM, SM, DMSM and TMSM methods for the NofI.

| Test function | MAGD | MSM | SM | AGD | TMSM | DMSM |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Perturbed Quadratic | 352325 | 34828 | 59908 | 353897 | 35697 | 28487 |
| Raydan 1 | 58504 | 26046 | 14918 | 22620 | 9801 | 17594 |
| Diagonal 3 | 119719 | 7030 | 12827 | 120416 | 8372 | 6409 |
| Generalized Tridiagonal 1 | 647 | 346 | 325 | 670 | 342 | 348 |
| Extended Tridiagonal 1 | 692219 | 1370 | 4206 | 3564 | 907 | 760 |
| Extended TET | 455 | 156 | 156 | 443 | 156 | 156 |
| Diagonal 4 | 8084 | 96 | 96 | 120 | 96 | 96 |
| Diagonal 5 | 48 | 72 | 72 | 48 | 72 | 72 |
| Extended Himmelblau | 302 | 260 | 196 | 396 | 288 | 294 |
| Perturbed quadratic diagonal | 1060824 | 37454 | 44903 | 2542050 | 31031 | 37331 |
| Quadratic QF1 | 362896 | 36169 | 62927 | 366183 | 39619 | 26585 |
| Extended quadratic penalty QP1 | 229 | 369 | 271 | 210 | 303 | 362 |
| Extended quadratic penalty QP2 | 356357 | 1674 | 3489 | 395887 | 2047 | 1908 |
| Quadratic QF2 | 71647 | 32727 | 64076 | 100286 | 39452 | 28651 |
| Extended quadratic exponential EP1 | 67 | 100 | 73 | 48 | 107 | 107 |
| Extended Tridiagonal 2 | 1665 | 659 | 543 | 1657 | 528 | 615 |
| ARWHEAD (CUTE) | 12834 | 430 | 270 | 5667 | 304 | 281 |
| Almost Perturbed Quadratic | 354369 | 33652 | 60789 | 356094 | 35755 | 26274 |
| LIARWHD (CUTE) | 925138 | 3029 | 18691 | 1054019 | 1340 | 3543 |
| ENGVAL1 (CUTE) | 822 | 461 | 375 | 743 | 418 | 482 |
| QUARTC (CUTE) | 177 | 217 | 290 | 171 | 289 | 275 |
| Generalized Quartic | 229 | 181 | 189 | 187 | 197 | 195 |
| Full Hessian FH3 | 63 | 63 | 63 | 45 | 63 | 63 |
| Diagonal 9 | 325609 | 10540 | 13619 | 329768 | 10219 | 11229 |

shows the performance CPUT.


Fig. 4.3: Performance profiles based on the NofI (left) and NofFE (right).

Table 4.6: Numerical results of the AGD, MAGD, MSM, SM, DMSM and TMSM methods for the NofFE.

| Test function | MAGD | MSM | SM | AGD | TMSM | DMSM |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Perturbed Quadratic | 13855459 | 200106 | 337910 | 13916515 | 423496 | 260678 |
| Raydan 1 | 1282162 | 311260 | 81412 | 431804 | 124905 | 280011 |
| Diagonal 3 | 4244404 | 38158 | 69906 | 4264718 | 95962 | 54865 |
| Generalized Tridiagonal 1 | 9057 | 1191 | 1094 | 9334 | 2408 | 2153 |
| Extended Tridiagonal 1 | 2077341 | 10989 | 35621 | 14292 | 13562 | 6800 |
| Extended TET | 4130 | 528 | 528 | 3794 | 1080 | 828 |
| Diagonal 4 | 133440 | 636 | 636 | 1332 | 1284 | 996 |
| Diagonal 5 | 108 | 156 | 156 | 108 | 300 | 228 |
| Extended Himmelblau | 5192 | 976 | 668 | 6897 | 2136 | 2418 |
| Perturbed quadratic diagonal | 38728371 | 341299 | 460028 | 94921578 | 619938 | 529154 |
| Quadratic QF1 | 13192789 | 208286 | 352975 | 13310016 | 472273 | 243573 |
| Extended quadratic penalty QP1 | 2939 | 2196 | 2326 | 2613 | 5073 | 3895 |
| Extended quadratic penalty QP2 | 8846145 | 11491 | 25905 | 9852040 | 29847 | 21345 |
| Quadratic QF2 | 2810965 | 183142 | 353935 | 3989239 | 444580 | 257674 |
| Extended quadratic exponential EP1 | 1513 | 894 | 661 | 990 | 2083 | 1617 |
| Extended Tridiagonal 2 | 9613 | 2866 | 2728 | 8166 | 4446 | 4456 |
| ARWHEAD (CUTE) | 468970 | 5322 | 3919 | 214284 | 9038 | 6761 |
| Almost Perturbed Quadratic | 13936462 | 194876 | 338797 | 14003318 | 424470 | 237534 |
| LIARWHD (CUTE) | 41619197 | 27974 | 180457 | 47476667 | 22254 | 53306 |
| ENGVAL1 (CUTE) | 8332 | 2285 | 2702 | 6882 | 6064 | 4442 |
| QUARTC (CUTE) | 414 | 494 | 640 | 402 | 1264 | 909 |
| Generalized Quartic | 1244 | 493 | 507 | 849 | 1043 | 798 |
| Full Hessian FH3 | 1955 | 566 | 631 | 1352 | 1152 | 957 |
| Diagonal 9 | 12984028 | 68189 | 89287 | 13144711 | 131327 | 125119 |



Fig. 4.4: Performance profiles arising from CPUT.

Table 4.7: Numerical results of the AGD, MAGD, MSM, SM, DMSM and TMSM methods for the CPUT.

| Test function | $M A G D$ | $M S M$ | $S M$ | $A G D$ | $T M S M$ | $D M S M$ |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Perturbed Quadratic | 6049.531 | 116.281 | 185.641 | 6756.047 | 219.328 | 134.781 |
| Raydan 1 | 334.266 | 31.906 | 36.078 | 158.359 | 44.828 | 66.484 |
| Diagonal 3 | 6401.969 | 52.609 | 102.875 | 5527.844 | 129.734 | 96.688 |
| Generalized Tridiagonal 1 | 7.781 | 1.469 | 1.203 | 11.344 | 2.969 | 2.969 |
| Extended Tridiagonal 1 | 8853.172 | 29.047 | 90.281 | 55.891 | 25.672 | 12.609 |
| Extended TET | 2.766 | 0.516 | 0.594 | 3.219 | 1.234 | 0.938 |
| Diagonal 4 | 16.172 | 0.203 | 0.141 | 0.781 | 0.344 | 0.172 |
| Diagonal 5 | 0.313 | 0.344 | 0.328 | 0.391 | 0.594 | 0.516 |
| Extended Himmelblau | 1.031 | 0.297 | 0.188 | 1.953 | 0.688 | 0.875 |
| Perturbed quadratic diagonal | 22820.172 | 139.625 | 185.266 | 44978.750 | 263.953 | 220.719 |
| Quadratic QF1 | 6846.453 | 81.531 | 138.172 | 12602.563 | 173.953 | 91.047 |
| Extended quadratic penalty QP1 | 1.063 | 1.000 | 0.797 | 1.266 | 2.781 | 1.813 |
| Extended quadratic penalty QP2 | 1872.797 | 3.516 | 6.547 | 3558.734 | 8.750 | 5.906 |
| Quadratic QF2 | 768.563 | 73.438 | 132.703 | 1582.766 | 169.266 | 98.141 |
| Extended quadratic exponential EP1 | 0.844 | 0.688 | 0.438 | 0.750 | 1.000 | 0.859 |
| Extended Tridiagonal 2 | 2.531 | 1.047 | 1.031 | 3.719 | 1.828 | 1.922 |
| ARWHEAD (CUTE) | 138.000 | 1.969 | 1.359 | 95.641 | 2.813 | 2.625 |
| Almost Perturbed Quadratic | 7086.563 | 73.047 | 133.516 | 13337.125 | 158.156 | 92.578 |
| LIARWHD (CUTE) | 15372.625 | 9.250 | 82.016 | 27221.516 | 5.250 | 17.406 |
| ENGVAL1 (CUTE) | 2.641 | 1.047 | 1.188 | 3.906 | 2.578 | 2.391 |
| QUARTC (CUTE) | 2.078 | 1.844 | 2.313 | 2.469 | 4.625 | 3.203 |
| Generalized Quartic | 0.500 | 0.281 | 0.188 | 0.797 | 0.422 | 0.500 |
| Full Hessian FH3 | 1.188 | 0.391 | 0.391 | 1.438 | 1.063 | 0.891 |
| Diagonal 9 | 6662.984 | 43.609 | 38.672 | 6353.172 | 61.984 | 114.703 |

Table 4.8: Average numerical results in the second numerical experiment.

| Average performances | $M A G D$ | $M S M$ | $S M$ | $A G D$ | TMSM | $D M S M$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| Number of iterations | 196051.21 | 9497.04 | 15136.33 | 235632.88 | 9058.46 | 8004.88 |
| No. of fun.evaluation | 6426009.58 | 67265.54 | 97642.88 | 8982579.21 | 118332.71 | 87521.54 |
| CPU time (sec) | 3468.58 | 27.71 | 47.58 | 5094.18 | 53.49 | 40.45 |

In accordance with obtained numerical data generated in the second numerical experiment, we can give an answer to the question, that independently of the choice of parameter values in the second and third backtracking line search, the DMSM iterations has the best results in relation to NofI. Also, if we compare the average results obtained in Tables 4.4 and 4.8, we can see that there is a slight percentage decrease in the average numerical results of the NofFE and CPUT, the DMSM method compared to the MSM method in the second numerical experiment.

## 5. Conclusion

Multiple usage of the backtracking line search in the modified SM (MSM)
method lead to two improvements of the MSM scheme, denoted as the TMSM and DMSM methods. Proposed iterations are investigated both theoretically and numerically. The linear convergence of the defined model is proved for UC and for a subset of SCQ functions. Numerical experiments confirm that the derived TMSM and DMSM methods outperform the SM, AGD, MAGD and the MSM with respect to the number of iterations. Numerical values arranged in Tables 4.1-4.8 confirm the better performance of presented accelerated gradient descent method. Finally, the obtained TMSM and DMSM methods can be used as a motivation for different possibilities of deriving new efficient schemes for unconstrained optimization.

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Branislav Ivanov
University of Belgrade, Technical Faculty in Bor
Department of Management
Vojske Jugoslavije 12, 19210 Bor, Serbia
ivanov.branislav@gmail.com

Bilall I. Shaini
State University of Tetova
Rr. e Ilindenit, p.n., Tetovo
R. Macedonia
bilall.shaini@unite.edu.mk

Predrag S. Stanimirović
University of Niš, Faculty of Sciences and Mathematics
Department of Computer Science
Višegradska 33, 18000 Niš, Serbia
pecko@pmf.ni.ac.rs

# SOLITARY WAVE SOLUTIONS FOR SPACE-TIME FRACTIONAL COUPLED INTEGRABLE DISPERSIONLESS SYSTEM VIA GENERALIZED KUDRYASHOV METHOD 

Ahmed A. Gaber and Hijaz Ahmad

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Abstract. In this article, space-time fractional coupled integrable dispersionless system has been considered, and we have used fractional derivative in the sense of modified Riemann-Liouville. The fractional system has been reduced to an ordinary by fractional transformation and the generalized Kudryashov method is applied to obtain exact solutions. We also testify performance as well as the precision of the applied method by means of numerical tests for obtaining solutions. The obtained results have been graphically presented to show the properties of the solutions.
Keywords. integrable dispersionless system; fractional derivative; differential system.

## 1. Introduction

In recent years, fractional differential equations have gained much attention from researchers due to thier numerous applications in many fields of sciences and engineering. These equations are widely used to describe various phenomena in many fields such as the fluid flow, electro chemistry, scattering theory, transport theory, probability, elasticity, control theory, potential theory, signal processing, image processing, diffusion theory, kinetic theory, systems identification, biology and other areas [1, 2]. The first application of fractional calculus was introduced by Abel [3] in the solution of an integral equation that was arisen in the formulation of the tautochronous problem. This problem deals with the determination of the shape of a frictionless plane curve through the origin in a vertical plane along with a particle of mass m can fall in a time that is independent of the starting position [4].

Travelling wave methods have an important role to obtain solutions that are described and explained these natural phenomena. Most famous of by these effective methods are ( $\mathrm{G}^{\prime} / \mathrm{G}$ )-expansion method $[5-6]$, variational iteration algorithm-I

[^6][7-8], Exp-function method [9], fractional iteration algorithm [10-11], Generalization of He's Exp-Function Method [12], reproducing Kernal method [13], a new extended Auxiliary equation method [14], variational iteration algorithm-II [ $15-16$ ] and Modified Kudryashov method [17-19]. In this paper, we use generalized Kudryashov method for finding the exact solutions of space-time fractional coupled integrable dispersionless system.

## 2. Properties of fractional derivatives

In this paper, we consider the most common definition named in modified RiemannLiouville derivative which is defined [20-26]

$$
D_{t}^{\gamma} u(t)=\left\{\begin{array}{cc}
\frac{1}{\Gamma(-\gamma)} \int_{0}^{t}(t-\tau)^{-\gamma-1}(u(\tau)-u(t)) d \tau & , \gamma<0  \tag{1}\\
\frac{1}{\Gamma(-\gamma)} \frac{d}{d t} \int_{0}^{t}(t-\tau)^{-\gamma-1}(u(\tau)-u(t)) d \tau, & 0<\gamma \leq 1 \\
\left(u^{(n-1)}(\tau)\right)^{(\gamma-n-1)}, & n-1<\gamma \leq n, n \geq 2
\end{array}\right.
$$

where $u: R \rightarrow R, t \rightarrow u(t)$, denotes a continuous function.
Property 1,

$$
\begin{equation*}
D_{t}^{\gamma} t^{r}=\frac{\Gamma(1+r)}{\Gamma(1+r-\gamma)} t^{r-\gamma}, r>0 \tag{2}
\end{equation*}
$$

Property 2,

$$
\begin{equation*}
D_{t}^{\gamma}(u(t) g(t))=g(t) D_{t}^{\gamma} u(t)+u(t) D_{t}^{\gamma} g(t) \tag{3}
\end{equation*}
$$

Property 3,

$$
\begin{equation*}
D_{t}^{\alpha} u(g(t))=\frac{d u(g(t))}{d g(t)} D_{t}^{\alpha} g(t) \tag{4}
\end{equation*}
$$

## 3. Description of the method for FDEs

Consider a given nonlinear wave equation

$$
\begin{equation*}
N\left(u, D_{t}^{\alpha} u, D_{x}^{\alpha} u, D_{x}^{2 \alpha} u_{x x}, D_{t}^{2 \alpha} u, D_{t}^{\alpha} D_{x}^{\alpha} u, \ldots\right)=0 \tag{5}
\end{equation*}
$$

we seek its wave solutions

$$
\begin{equation*}
u=U(\eta), \quad \eta=\frac{h_{i} x_{i}^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}, \quad i=1,2, \ldots \tag{6}
\end{equation*}
$$

Consequently, Eq. (5) is reduced to the ordinary differential equation (ODE) by transformation:

$$
\begin{equation*}
U\left(u, g u^{\prime}, h u^{\prime}, g^{2} u^{\prime \prime}, h^{2} u^{\prime \prime}, \ldots\right)=0 \tag{7}
\end{equation*}
$$

The generalized Kudryashov method (GKM) is based on the assumption that the travelling wave solutions can be expressed in the following form

$$
\begin{equation*}
u(\eta)=\sum_{i=0}^{m} \frac{a_{i}}{(1+\phi(\eta))^{i}} \tag{8}
\end{equation*}
$$

where $m$ is positive integer which are unknown to be further determined, $a_{i}$ are unknown constants. In addition, $\phi(\eta)$ satisfies Riccati equation

$$
\begin{equation*}
\phi^{\prime}(\eta)=A+B \phi(\eta)+C \phi^{2}(\eta) \tag{9}
\end{equation*}
$$

We obtained a type of solutions of Eq. (9)
Family 1: $A$ and $B$ are free constants, $C \neq 0$

$$
\phi(\eta)=\frac{-B+\sqrt{4 A C-B^{2}} \tan \left(\frac{1}{2}\left(\sqrt{4 A C-B^{2}}\left(\eta+d_{0}\right)\right)\right)}{2 C}
$$

Family 2: $A=0, B \neq 0$, and $C$ is a free constant

$$
\phi(\eta)=\frac{-B \exp \left(B \eta+B d_{0}\right)}{C \exp \left(B \eta+B d_{0}\right)-1}
$$

Family 3: $A$ is free constant, $B \neq 0$, and $C=0$

$$
\phi(\eta)=\frac{-A}{B}+\frac{1}{B} \exp (B \eta)
$$

Family 4: $A=0, B=-1$ and $\mathrm{C}=-1$

$$
\phi(\eta)=\frac{-d_{0}}{\exp (\eta)+d_{0}}
$$

## 4. Space-time fractional coupled Integrable Dispersionless system

We consider the space-time fractional coupled Integrable Dispersionless (CID) system

$$
\begin{aligned}
\frac{\partial^{2 \alpha} u}{\partial t^{\alpha} \partial x^{\alpha}}+\frac{\partial^{\alpha}}{\partial x^{\alpha}}(v w) & =0 \\
\frac{\partial^{2 \alpha} v}{\partial t^{\alpha} \partial x^{\alpha}}-2 v \frac{\partial^{\alpha} u}{\partial x^{\alpha}} & =0
\end{aligned}
$$

$$
\begin{equation*}
\frac{\partial^{2 \alpha} w}{\partial t^{\alpha} \partial x^{\alpha}}-2 w \frac{\partial^{\alpha} u}{\partial x^{\alpha}}=0 \tag{10}
\end{equation*}
$$

where $u, v$ and $w$ are all functions of $x$ and $t$. Eqs. (10) describes the currentfed string within an external magnetic field [27,28]. This equations wase presented and solved by the inverse scattering method [29], the exp-function method [30] and residue harmonic balance [31].

We perform the transformation $\eta=\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}$, Eq. (10) can be reduced into an ODEs

$$
\begin{aligned}
g h U^{\prime \prime}+h\left(V W^{\prime}+W V^{\prime}\right) & =0 \\
g h V^{\prime}-2 h V U^{\prime} & =0
\end{aligned}
$$

$$
\begin{equation*}
g h W^{\prime}-2 h W U^{\prime}=0 \tag{11}
\end{equation*}
$$

where $U^{\prime}=\frac{\partial U}{\partial \eta}$.
We can freely know that the solution does not depend on the balancing the highest order linear and nonlinear terms [32]. For simplicity, we set $i=2$, we have:

$$
\begin{align*}
U(\eta) & =a_{0}+\frac{a_{1}}{1+\phi(\eta)}+\frac{a_{2}}{(1+\phi(\eta))^{2}} \\
V(\eta) & =b_{0}+\frac{b_{1}}{1+\phi(\eta)}+\frac{b_{2}}{(1+\phi(\eta))^{2}} \\
W(\eta) & =r_{0}+\frac{r_{1}}{1+\phi(\eta)}+\frac{r_{2}}{(1+\phi(\eta))^{2}} \tag{12}
\end{align*}
$$

Substituting Eq. (12) into Eq. (11), equating to zero the coefficients of all powers of $\phi(\eta)$ yields a set of algebraic equations for $\mathrm{a}_{i}, \mathrm{~b}_{i}, \mathrm{r}_{i}$.

$$
\begin{aligned}
& h\left(6 B C r_{0} a_{1}+2 A r_{0} a_{1}+3 g r_{1} A B+4 B r_{0} a_{2}+g r_{1} B^{2} C+2 B C r_{1} a_{1}+4 g r_{2} B^{2}\right)=0, \\
& h\left(2 A b_{0} a_{1}+6 B b_{0} a_{1}+4 g b_{2} B^{2}+3 g b_{1} A B+g b_{1} B^{2}+4 B C b_{0} a_{2}+2 B b_{1} a_{1}\right)=0, \\
& h\left(2 B r_{0} a_{1}+g r_{1} B^{2}\right)=0, h\left(g b_{1} B^{2}+2 B b_{0} a_{1}\right)=0, h\left(g a_{1} B^{2}-B b_{0} r_{1}-B b_{1} r_{0}\right)=0, \\
& h\left(g a_{1} B^{2}-A b_{1} r_{0}+3 g a_{1} A B-2 B C b_{0} r_{2}-A b r_{1}-2 B b_{2} r-2 B b_{1} r_{1}-3 B C b_{0} r_{1}\right. \\
& \left.+4 g a_{2} B^{2}-3 B b_{1} r_{0}\right)=0, \\
& h\left(2 g r_{1} A^{2}+6 g r_{2} A^{2}+2 A r_{0} a_{1}+4 A r_{0} a_{2}+2 A r_{1} C a_{1}+4 A r_{1} a_{2}+2 A C r_{2} a_{1}\right. \\
& \left.+4 A r_{2} a_{2}-2 g r_{2} A B-g r_{1} A B\right)=0, \\
& h\left(4 A b_{1} a_{2}+2 A b_{2} a_{1}+4 A b_{2} a_{2}+2 A b_{0} a_{1}+4 A b_{0} a_{2}+6 g b_{2} A^{2}-g b_{1} A B\right. \\
& \left.-2 g b_{2} A B+2 A b_{1} a_{1}+2 g b_{1} A^{2}\right)=0, \\
& h\left(-2 A b_{1} r_{1}+2 g a_{1} A^{2}+6 g a_{2} A^{2}-A C b_{1} r_{0}-4 A b_{2} r_{2}-3 A C b_{1} r_{2}-2 A b_{2} r_{0}\right.
\end{aligned}
$$

$$
\begin{equation*}
\left.-3 A b_{2} r_{1}-A b_{0} r_{1}-g a_{1} A B-2 A b_{0} r_{2}-2 g a_{2} A B\right)=0, \tag{13}
\end{equation*}
$$

Solving the system of algebraic equations with the help of Maple, we obtain the solutions organized in the following cases:

Case (1)

$$
\begin{align*}
& a_{1}=-g A-g C+g B, b_{0}=\frac{-1}{4 r_{0}} g^{2}\left(4 C^{2}-4 B C+B^{2}\right), \\
& b_{1}=\frac{-1}{2 r_{0}} g^{2}\left(-2 A C-2 C^{2}+3 B C+A B-B^{2}\right), \\
& r_{1}=\frac{2 r_{0}(A+C-B)}{-2 C+B}, a_{0} \text { is arbitrary, } b_{2}=r_{2}=a_{2}=0 . \tag{14}
\end{align*}
$$

Substituting these results into (11) and with the aid of families 1-4, we obtain the following multiple soliton-like and periodic solutions for space-time fractional CID system

$$
\begin{align*}
& u(x, t)=a_{0}+\frac{-g A-g C+g B}{1+\frac{-B+\sqrt{4 A C-B^{2}} \tan \left(\frac{1}{2}\left(\sqrt{4 A C-B^{2}}\left(\eta+d_{0}\right)\right)\right)}{2 C}}, \\
& v(x, t)=\frac{-1}{4 r_{0}} g^{2}\left(4 C^{2}-4 B C+B^{2}\right)+\frac{-2 C g^{2}\left(-2 A C-2 C^{2}+3 B C+A B-B^{2}\right)}{2 r_{0}\left(2 C-B+\sqrt{4 A C-B^{2}} \tan \left(\frac{1}{2}\left(\sqrt{4 A C-B^{2}}\left(\eta+d_{0}\right)\right)\right)\right)}, \\
& w(x, t)=r_{0}+\frac{4 C r_{0}(A+C-B)}{(-2 C+B)\left(2 C+-B+\sqrt{4 A C-B^{2}} \tan \left(\frac{1}{2}\left(\sqrt{4 A C-B^{2}}\left(\eta+d_{0}\right)\right)\right)\right)}, \tag{15}
\end{align*}
$$

where $\eta=\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}$.

## Case (2)

$$
\begin{gather*}
a_{1}=g C, b_{1}=\frac{g^{2} C^{2}}{r_{1}}, B=2 C, r_{1} \text { is arbitrary } \\
r_{0}=b_{0}=a_{2}=b_{2}=r_{2}=A=0 \tag{16}
\end{gather*}
$$

Substituting these results into (11) and with the aid of families 1-4, we obtain the following multiple soliton-like and periodic solutions for space-time fractional CID system

$$
\begin{aligned}
& u(x, t)=a_{0}-\frac{g C\left(C \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)+B d_{0}\right)-1\right)}{C \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)+B d_{0}\right)+1}, \\
& v(x, t)=b_{0}-\frac{g^{2} C^{2}\left(C \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)+B d_{0}\right)-1\right)}{r_{1}\left(C \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)+B d_{0}\right)+1\right)}
\end{aligned}
$$

$$
\begin{equation*}
w(x, t)=r_{0}-\frac{r 1\left(C \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)+B d_{0}\right)-1\right)}{C \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)+B d_{0}\right)+1} . \tag{17}
\end{equation*}
$$

## Case (3)

$$
\begin{gather*}
a_{1}=g(B-A), b_{0}=\frac{-g^{2} B^{2}}{4 r_{0}}, r_{1}=\frac{-2 r_{0}(A-B)}{B}, a_{0} \text { is arbitrary } \\
b_{1}=\frac{-g^{2} B(A-B)}{2 r_{0}}, a_{2}=b_{2}=r_{2}=0 \tag{18}
\end{gather*}
$$

Substituting these results into (11) and with the aid of families 1-4, we obtain the following multiple soliton-like and periodic solutions for space-time fractional CID system

$$
\begin{align*}
u(x, t) & =a_{0}+\frac{g(B-A)}{1-\frac{A}{B}+\frac{1}{B} \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)\right)} \\
v(x, t) & =\frac{-g^{2} B^{2}}{4 r_{0}}-\frac{g^{2} B(A-B)}{2 r_{0}\left(1-\frac{A}{B}+\frac{1}{B} \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)\right)\right.}, \\
w(x, t) & =r_{0}-\frac{-2 r_{0}(A-B)}{B\left(1-\frac{A}{B}+\frac{1}{B} \exp \left(B\left(\frac{h x^{\alpha}}{\Gamma(1+\alpha)}+\frac{g t^{\alpha}}{\Gamma(1+\alpha)}\right)\right)\right.} . \tag{19}
\end{align*}
$$

## 5. Discussion

In this section, we discuss the physical explanations of the obtained solutions. Note that, the plots of the solutions (15), (17), and (19) are presented in figures 1 to 6 at specific values of the free constants. It appears that the solutions of (15), (17) and (19) depend on the sign of the magnitude $4 A C-B^{2}$. In the case of $4 A C-B^{2}>0$, the solution (15) is expressed in terms of the trigonometric tan function and hence an anti-kink wave is produced as shown by Fig. 7.1(a). Similarly the solution (19) in Fig. 7.3(a). On the other side, $4 \mathrm{AC}-\mathrm{B}^{2}<0$, the solution (17) can be expressed in terms of the hyperbolic tan function and accordingly a kink wave is resulted as displayed in Fig. 7.2(b).

On the other hand, the solutions are affected by the fractional derivatives $\alpha$, in Fig. 7.1(b) the anti-kink wave increases with increasing of $\alpha$ but the reverse effect is observed a little near off the plate $(x>0.5)$. Similarly the solution (19) in Fig. 7.1(b). Finally, Fig. 7.2(b) describes the u-solution in (17), the kink wave increases with increasing of $\alpha$ at $0<x \leq 1$ and then it stabilizes with different value of $\alpha$ at $x>1$. Accordingly, this method is capable of producing a different types of wave solutions for partial differential equations

## 6. Conclusion

In the present paper, GKM has been successfully employed to obtain the exact solution of space-time fractional coupled integrable dispersionless system. New travelling wave technique is applied to search for the exact solitary solutions. The main advantage of the proposed method over the others is the fact that it can be applied to a wide class of nonlinear evolution equations. The modified Kudryashov [16] is special case of this technique (family 3 , take $\mathrm{A}=0$ and $\mathrm{B}=1$ ). Finally, the obtained results have been graphically presented to show the properties of the obtained solutions.

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Fig. 7.1: (a) Anti-Kink wave solution of Eq. (17) where $a_{0}=a_{1}=g=A=B=C$ $=1, d_{0}=0, h=0.1$ and $\alpha=\frac{1}{2}$
(b) Anti-Kink wave solution of Eq. (17) where $\mathrm{t}=0, a_{0}=a_{1}=g=A=B=C=1, d_{0}=0$ and $h=0.1$


Fig. 7.2: (a) Kink wave solution of Eq. (15) where $a_{0}=1, C=2, g=-1, d_{0}=0, h=1$ and $\alpha=\frac{3}{4}$.
(b) Kink wave solution of Eq. (15) where $\mathrm{t}=0, a_{0}=1, C=1, g=-1, d_{0}=0$ and $h=1$.


Fig. 7.3: (a) (5) Anti-Kink wave solution of Eq. (19) where $a_{0}=\mathrm{r}_{0}=1, b_{1}=-0.1, g=$ $B=1, h=0.1$ and $\alpha=\frac{1}{2}$.
(b) Anti-Kink wave solution of Eq. (19) where $a_{0}=\mathrm{r}_{0}=1, b_{1}=-0.1, g=B=1, h=0.1$

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Ahmed A. Gaber<br>College of Science and Humanities at Howtat Sudair<br>Department of Mathematics, Majmaah University,<br>Majmaah 11952, Saudi Arabia.<br>Faculty of Education, Department of Mathematics, Ain Shams University, Heliopolis, Roxy, Egypt.<br>a.gaber@mu.edu.sa and aagaber6@gmail.com

Hijaz Ahmad
Section of Mathematics, International Telematic University Uninettuno,
Corso Vittorio Emanuele II, 39, 00186 Roma, Italy.
Department of Basic Sciences, University of Engineering and Technology Peshawar, 25000, Pakistan
hijaz555@gmail.com

# AN IDENTITY-BASED ENCRYPTION SCHEME USING ISOGENY OF ELLIPTIC CURVES 

Mojtaba Bahramian* and Elham Hajirezaei

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#### Abstract

Identity-Based Encryption is a public-key cryptosystem that uses the receiver identifier information such as email address, IP address, name, etc, to compute a public and a private key in a cryptosystem and encrypt a message. A message receiver can obtain the secret key corresponding with his privacy information from private key generator and he can decrypt the ciphertext. In this paper, we review Boneh-Franklin's scheme and use a bilinear map and Weil pairing's properties to propose an identitybased cryptography scheme based on isogeny of elliptic curves.


Keywords: Identity-based encryption; elliptic curves; isogeny of elliptic curves.

## 1. Introduction

Public key encryption (PKE), involves two distinct keys, public key, and private key. The public key can be widely distributed without compromising its corresponding private key. Identity-Based Encryption (IBE) is a public-key encryption scheme in which the public key can be an arbitrary string. Identity-based encryption is a cryptographic scheme, which enables any pair of users to communicate securely without exchanging secret or public keys. Actually by the identity-based scheme, if you know somebody's name or email address you can send him a message which only he can read. This issue has now been particularly attended by cryptographic researchers and so far, many cryptography schemes are based on it has been presented.

The basic identity scheme was first proposed by Shamir [11] in 1984. The scheme is specified by four phases:

1. Setup: In this phase, general system parameters and master-key are created.
2. Extraction: In this algorithm, the private key associated with an arbitrary public key string $I D \in\{0,1\}^{*}$ is created by using the master-key.

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3. Encryption: A message is encrypted using the public key $I D$.
4. Decryption: An encrypted message is decrypted having the corresponding private key.

When the sender, Alice, sends an e-mail to the receiver, Bob, at bob@email.com, she simply encrypts her message having the public key string "bob@email.com". In this method, we need a trusted third party known as "Private Key Generator" (PKG), which computes a master private key and a public key. The PKG has a privileged position by knowing some secret information that enables it to compute the private keys for all the users in the system. Thus, when Bob receives the encrypted message by his e-mail, he contacts to the PKG, authenticates himself to it in the same way, then he obtains his private key from the PKG, and he can read his e-mail $[1,6]$. The problem of constructing an IBE was an open problem for many years. Finally, Boneh and Franklin [1] proposed an IBE scheme using bilinear maps in 2001. Soon after Boneh and Franklin's announcement, it was detected that Clifford Cocks, had designed a simple IBE years earlier.

Boneh and Franklin presented a functional IBE scheme in which the performance of their approach is similar to the performance of ElGamal encryption in $\mathbb{F}_{q}^{*}$, and the security of their scheme is based on the Computational Diffie-Hellman (CDH) hypothesis on elliptic curves.

In this paper, we propose an identity-based encryption scheme based on the isogenies between elliptic curves. The security of our scheme is based on the hardness of the isogeny problem that is finding an isogeny between two given isogenous elliptic curves. In our proposed scheme we use the endomorphism ring of an ordinary elliptic curve $E$, $(\operatorname{End}(\mathrm{E}))$, and some its properties such as the commutativity of $\operatorname{End}(E)$.
Basic Concepts of IBE. As mentioned earlier, in the IBE scheme Alice can use the receiver's identifier information which is presented by any string, such as email address or IP address, even a digital image [10], to encrypt a message. Bob obtains a private key corresponding to his identifier information from the trusted third party, then he can decrypt the ciphertext (Fig. 1.1).

Universally an identity-based encryption scheme is specified by four randomized algorithms:

1. Setup: First, the PKG creates a public key $p k_{P K G}$ and a master private key $s k_{P K G}$, then he publishes $p k_{P K G}$ as a public key.
2. Extraction: Bob authenticates himself to the PKG and receives his private key $s k_{B o b}$ corresponding to his identity, $I D_{B o b}$.
3. Encryption: Alice encrypts her message, $M$ to the ciphertext $C$ using $I D_{B o b}$ and $p k_{P K G}$.
4. Decryption: Bob decrypts the ciphertext $C$, using his private key, $s k_{B o b}$ and reconstruct the message $M$.


Fig. 1.1: Identity-based encryption scheme

The rest of the paper is organized as follows: Section 2 contains a summary of some preliminaries on elliptic curves, isogenies, and basic properties of the Weil pairing. In section 3, we give a review of Boneh and Franklin's IBE scheme. Our proposed identity-based encryption scheme is given in Section 4. Finally, we dedicate the security analysis of our scheme in Section 5.

## 2. Preliminaries

In this section, we first briefly introduce elliptic curves, isogenies and Weil pairing (see [12, 15]).

### 2.1. Elliptic Curves

Elliptic Curve Cryptography (ECC) was introduced by Koblitz [5] and Miller [8] in 1985. They proposed completely different cryptographic use of elliptic curves. The main reason for the attractiveness of ECC is the fact that there is no subexponential algorithm known for solving the Discrete Logarithm Problem (DLP) on a properly chosen elliptic curve. We will refer to it later.

Definition 2.1. Let $K$ be a field of characteristic not equal to 2 and 3. An elliptic curve $E$ over $K$ is a curve given by a (short) Weierstrass equation of the form

$$
\begin{equation*}
y^{2}=x^{3}+A x+B \tag{2.1}
\end{equation*}
$$

where $A, B \in \bar{K}$, and its discriminant, $\Delta=-16\left(4 A^{3}+27 B^{2}\right)$ is nonzero. The $j$-invariant of the elliptic curve $E$ is defined by

$$
j=j(E)=1728 \frac{4 A^{3}}{4 A^{3}+27 B^{2}}
$$

furthermore, any elliptic curve $E$ can be determined by its $j$-invariant. In other words, two elliptic curves with the same $j$-invariant are isomorphic over $K$.

We say that the elliptic curve $E: y^{2}=x^{3}+A x+B$ is defined over $K$, where $A, B \in K$. For the elliptic curve $E$ defined over $K$, the set of $K$-rational points of $E$ is defined by

$$
E(K)=\left\{(x, y) \in K^{2}: y^{2}=x^{3}+A x+B\right\} \cup\{\mathcal{O}\}
$$

where, $\mathcal{O}$ is the point at infinity.
The set $E(K)$ forms an abelian additive group with identity element $\mathcal{O}$. Let $P=\left(x_{P}, y_{P}\right)$ and $Q=\left(x_{Q}, y_{Q}\right)$ be two points on the curve. The sum of $P$ and $Q$ is defined as $R=P+Q=\left(x_{R}, y_{R}\right)$ where,

1. If $x_{P} \neq x_{Q}$, then $x_{R}=m^{2}-x_{P}-x_{Q}$ and $y_{R}=m\left(x_{P}-x_{R}\right)-y_{P}$, where $m=\left(y_{Q}-y_{P}\right) /\left(x_{Q}-x_{P}\right)$.
2. If $x_{P}=x_{Q}$ and $y_{P} \neq y_{Q}$, then $R=\mathcal{O}$.
3. If $P=Q$ and $y_{P} \neq 0$, then $x_{R}=m^{2}-2 x_{P}$ and $y_{R}=m\left(x_{P}-x_{R}\right)-y_{P}$, where, $m=\left(3 x_{P}^{2}+A\right) / 2 y_{P}$.
4. If $P=Q$ and $y_{P}=0$, then $R=\mathcal{O}$.
5. If $Q=\mathcal{O}$, then $R=P$.

For the Weierstrass equation described by (2.1), if $P=(x, y)$, then $-P=(x,-y)$.
Suppose $E$ is an elliptic curve defined over a field $K$ and Let $n$ be a positive integer, the $n$-torsion subgroup of $E$ defined as follows

$$
E[n]=\{P \in E(\bar{K}) \mid n P=\mathcal{O}\}
$$

If the characteristic of $K$ does not divide $n$, or is zero, then $E[n] \cong \mathbb{Z}_{n} \times \mathbb{Z}_{n}$, and if the characteristic of $K$ is $p>0, n=p^{r} n^{\prime}$ with $p \nmid n$, then $E[n] \cong \mathbb{Z}_{n^{\prime}} \times \mathbb{Z}_{n}$ or $\mathbb{Z}_{n^{\prime}} \times \mathbb{Z}_{n^{\prime}}$. For the elliptic curve $E$ defined over the finite field $\mathbb{F}_{q}, q=p^{r}$ for some prime $p$, we say that $E$ is supersingular if $E[p]=\{\mathcal{O}\}$, and $E$ is called ordinary if $E[P] \cong \mathbb{Z}_{p}$.

Let the elliptic curve $E$ defined over the field $\mathbb{F}_{q}$. Then $E\left(\mathbb{F}_{q}\right) \cong Z_{n}$ for some integer $n \geq 1$, or $E\left(\mathbb{F}_{q}\right) \cong Z_{n_{1}} \times Z_{n_{2}}$ for some integers $n_{1}, n_{2} \geq 1$ with $n_{1}$ dividing $n_{2}$. By Hasse's theorem, for elliptic curve $E$ over the finite field $\mathbb{F}_{q}$, the order of $E$ satisfies $\left|q+1-\# E\left(\mathbb{F}_{q}\right)\right| \leq 2 \sqrt{q}$. The trace of the elliptic curve $E$ denoted by $a_{q}$, is $a_{q}=q+1-\# E(F q)$. The elliptic curve $E$ is supersingular if and only if $a_{q} \equiv 0$ $(\bmod p)$, it means that $\# E\left(\mathbb{F}_{q}\right) \equiv 1 \quad(\bmod p)$.
Discrete Logarithm Problem: Let $E$ be an elliptic curve defined over the finite field $\mathbb{F}_{q}, P \in E$ and $Q \in\langle P\rangle$. The Elliptic Curve Discrete Logarithm Problem (ECDLP) is the problem of finding integer $n$ such that $Q=n P$. It is Well-known that the fastest known algorithm to solve the ECDLP over an arbitrary curve is Pollard's rho method, which has exponential time complexity. [9].

### 2.2. Isogeny of Elliptic Curves

Definition 2.2. Let $K$ be a field and let $E_{1}$ and $E_{2}$ be two elliptic curves defined over $K$. An isogeny is a non-constant morphism $\varphi: E_{1}(\bar{K}) \rightarrow E_{2}(\bar{K})$ satisfying $\varphi\left(\mathcal{O}_{E_{1}}\right)=\mathcal{O}_{E_{2}}$. The isogeny $\varphi$ can be displayed by

$$
\varphi:(x, y) \rightarrow\left(\frac{p(x)}{q(x)}, \frac{r(x)}{s(x)} y\right)
$$

with polynomials $p(x), q(x), r(x)$ and $s(x)$ such that $p(x)$ and $q(x)$ do not have a common factor. The degree of isogeny $\varphi$ denoted $\operatorname{by} \operatorname{deg}(\varphi)$, is the maximum degree of the polynomials $p(x)$ and $q(x)$. Also, we define $\operatorname{deg}(\mathbf{0})=0$. The isogeny $\varphi$ is called separable, if $\operatorname{deg}(\varphi)=\# \operatorname{ker}(\varphi)$. We say that two elliptic curves $E_{1}$ and $E_{2}$ are $l$-isogenous when there exists a nonzero isogeny of degree $l$ from $E_{1}$ to $E_{2}$. If $\varphi: E_{1} \rightarrow E_{2}$ is an isogeny of degree $l$, then the dual of $\varphi$ denoted by $\hat{\varphi}$, is a unique isogeny from $E_{2}$ to $E_{1}$ of the same degree $l$, such that $\hat{\varphi} \circ \varphi=[l]_{E_{1}}$, the multiplication by $l$ map on $E_{1}$ and also, $\varphi \circ \hat{\varphi}=[l]_{E_{2}}$. By Tate's theorem [9], two elliptic curves $E_{1}$ and $E_{2}$ are isogenous over the finite field $\mathbb{F}_{q}$, if and only if $\# E_{1}\left(\mathbb{F}_{q}\right)=\# E_{2}\left(\mathbb{F}_{q}\right)$. We denote the set of isogenies from $E_{1}$ to $E_{2}$ by $\operatorname{Hom}\left(E_{1}, E_{2}\right)$. The sum of two isogenies $\varphi$ and $\psi$ is defined by $(\varphi+\psi)(P)=\varphi(P)+\psi(P)$, for each $P \in E$. It implies that $\varphi+\psi$ is an isogeny, and thus $\operatorname{Hom}\left(E_{1}, E_{2}\right)$ is a group. If $E_{1}=E_{2}$, then we can also compose isogenies. If $E$ is an elliptic curve, we let $\operatorname{End}(E)=\operatorname{Hom}(E, E)$ be the ring whose addition law is as given above and whose multiplication is composition, $(\varphi \psi)(P)=\varphi(\psi(P))$. The ring $\operatorname{End}(E)$ is called the endomorphism ring of $E$. The Frobenius endomorphism $\tau_{q}$ is defined by $\tau_{q}(x, y)=\left(x^{q}, y^{q}\right)$. It is an endomorphism of $E$ (see [15]).

### 2.3. Bilinear Map

Let $G_{1}$ be an additive group of order $r$ and $G_{2}$ be a multiplicative group of the same order. A function $e: G 1 \times G 1 \rightarrow G_{2}$ is said to be a bilinear pairing if the following properties hold

1. Bilinearity: for all $P, Q \in G_{1}$ and $a, b \in \mathbb{Z}_{r}^{*}, e(a P, b Q)=e(P, Q)^{a b}$.
2. Non-degeneracy: there exist $P, Q \in G_{1}$ such that $e(P, Q) \neq 1$.
3. Computability: for all $P, Q \in G_{1}$, there exists an efficient algorithm to compute $e(P, Q)$.

As we will say in section 2.4, the example of an efficiently computable non-degenerate the bilinear map is the Weil pairing.

### 2.4. Weil Pairing

As already mentioned if $E$ be an elliptic curve over a field $K$ and let $n$ be an integer not divisible by the characteristic of $K$, Then $E[n] \cong \mathbb{Z}_{n} \times \mathbb{Z}_{n}$. Let

$$
\mu_{n}=\left\{x \in \bar{K} \mid x^{n}=1\right\},
$$

be the group of $n$-th roots of unity in $K$. Since $n$ is not divided by the characteristic of $K$, the equation $x^{n}=1$ has no multiple roots so, it has $n$ distinct roots in $\bar{K}$, Therefore, $\mu_{n}$ is a cyclic group of order $n$. Any generator $\gamma$ of $\mu_{n}$ is called a primitive $n$th root of unity. This is equivalent to saying that $\gamma^{k}=1$ if and only if $k$ divided by $n$.

Definition 2.3. Let $E$ be an elliptic curve over a field $K$ and let $n$ be a positive integer not divisible by the characteristic of $K$. Then there is a pairing

$$
\begin{equation*}
e_{n}=E[n] \times E[n] \rightarrow \mu_{n} \tag{2.2}
\end{equation*}
$$

called the Weil Pairing. This concept satisfies the following properties:

1. $e_{n}$ is bilinear in each variable. This means that

$$
\begin{equation*}
e_{n}\left(S_{1}+S_{2}, T\right)=e_{n}\left(S_{1}, T\right) e_{n}\left(S_{2}, T\right) \tag{2.3}
\end{equation*}
$$

and

$$
\begin{equation*}
e_{n}\left(S, T_{1}+T_{2}\right)=e_{n}\left(S, T_{1}\right) e_{n}\left(S, T_{2}\right) \tag{2.4}
\end{equation*}
$$

for all $S, S_{1}, S_{2}, T, T_{1}, T_{2} \in E[n]$.
2. $e_{n}$ is nondegenerate in each variable. This means that if $e_{n}(S, T)=1$ for all $T \in E[n]$ then $S=\infty$ and also that if $e_{n}(S, T)=1$ for all $S \in E[n]$ then $T=\infty$.
3. $e_{n}(T, T)=1$ for all $T \in E[n]$.
4. $e_{n}(S, T)=e_{n}(T, S)^{-1}$ for all $S, T \in E[n]$.
5. $e_{n}(\sigma(S), \sigma(T))=\sigma\left(e_{n}(S, T)\right)$. For all automorphism $\sigma$ of $\bar{K}$ such that $\sigma$ is the identity map on the coefficient of $E$ (if $E$ is in Weiratrass form, this means that $\sigma(A)=A$ and $\sigma(B)=B)$.
6. $e_{n}\left(\alpha(S), \alpha(T)=e_{n}(S, T)^{\operatorname{deg}(\alpha)}\right.$ for all separable endomorphisms $\alpha$ of $E$.

If the coefficient of $E$ lie in a finite field $\mathbb{F}_{q}$, the statement also holds when $\alpha$ is the Frobenius endomorphism $\tau_{Q}$. (Actually, the statement holds for all endomorphism $\alpha$, separable or not.)

Now we say that the isogenies $\varphi$ and $\hat{\varphi}$ are dual (or adjoint) concerning the Weil pairing. Let $\varphi: E_{1} \rightarrow E_{2}$ be an isogeny of elliptic curves and let $\hat{\varphi}$ be its dual, and let $e_{n}$ be a Weil pairing. Then $e_{n}(\varphi(S), T)=e_{n}(S, \hat{\varphi}(T))$ for all $n$-torsion points $S \in E_{1}[n]$ and $T \in E_{2}[n]$ (see [7]).

## 3. Boneh-Franklin Scheme

Boneh and Franklin's Scheme can be built from any bilinear map ê : $G_{1} \times G_{1} \rightarrow G_{2}$ between two groups $G_{1}$ and $G_{2}$ as long as a variant of the computational DiffieHellman problem in $G_{1}$ is hard. They use the Weil pairing on elliptic curves as an example of such a map. They describe the scheme in four phases:

1. Setup: The PKG specifies an elliptic curve $E$ over $\mathbb{F}_{p}$. It Chooses an arbitrary $P \in E / \mathbb{F}_{p}$ of order $q$. The PKG also specifies two hash functions $H_{1}: \mathbb{F}_{p^{2}} \rightarrow$ $\{0,1\}^{n}$ and $H_{2}:\{0,1\}^{*} \rightarrow \mathbb{F}_{p}$. The $P K G$ picks a random $s \in Z_{q}^{*}$ as a master key and denoted it by $p k_{P K G}$. Then it computes a public key $p k_{P K G}=s P$. The PKG publishes $\left\{E, \mathbb{F}, P, H_{1}, H_{2}, p k_{P K G}\right\}$.
2. Extraction: Bob contacts the PKG to get his private key. The PKG first maps, Bob's identity, $I D_{\text {Bob }} \in\{0,1\}^{*}$ to a point $Q_{I D} \in E / \mathbb{F}_{p}$ of order $q$, then it computes $s k_{B o b}=s Q_{I D}$ where $Q_{I D}=H_{1}(I D)$ and $s$ is the master key.
3. Encryption: Alice encrypt her message $M \in\{0,1\}^{l}$ (where $l$ denotes the length of $M$ ). under the public key, $p k_{P K G}$ and $I D_{B o b}$ which is mapped to a point $Q_{I D} \in E / \mathbb{F}_{p}$ of order $q$. She computes $U=r P$ and $V=$ $H_{2}\left(\hat{e}\left(Q_{I D}, p k_{P K G}\right)^{r}\right) \oplus M$, where $r$ is chosen at random from $Z_{q}$ and $Q_{I D}=$ $H_{1}(I D)$. The resulting ciphertext $C=(U, V)$ is sent to Bob.
4. Decryption: Bob receives the ciphertext $C$, and checks it. If $U \in E / \mathbb{F}_{p}$ is not a point of order $q$ rejects the ciphertext. Otherwise, to decrypt $C$ using his private key, $s k_{B o b}$ and computes:

$$
\begin{equation*}
V \oplus H_{2}\left(\hat{e}\left(s k_{B o b}, U\right)\right)=M \tag{3.1}
\end{equation*}
$$

This completes the description as follows:

$$
\begin{aligned}
\hat{e}\left(s k_{B o b}, U\right) & =\hat{e}\left(s Q_{I D}, r P\right) \\
& =\hat{e}\left(Q_{I D}, P\right)^{s r} \\
& =\hat{e}\left(Q_{I D}, p k_{P K G}\right)^{r}
\end{aligned}
$$

Thus, applying decryption after encryption produces the original message $M$ as required.

## 4. Proposed Scheme

This section details our newly proposed identity-based encryption using isogeny of elliptic curves.

Let $\mathbb{F}_{q}$ be the field of order $q$, where $q$ is a power of a prime number $p$ and $n$ be a positive integer coprime to $p$. Let $E$ be an ordinary elliptic curve over $\mathbb{F}_{q}$, and let
$e_{n}: E[n] \times E[n] \rightarrow \mu_{n}$ be the Weil $e_{n}$-pairing. In our scheme, we use an algorithm $\mathcal{A}$ to convert a string $I D_{B o b} \in\{0,1\}^{*}$ to a point $Q_{I D} \in E$ of order $n$.

The phases in the proposed scheme are Setup phase, Extraction phase, Encryption phase and Decryption phase. The procedure of our scheme is described in detail as follows:

1. Setup: The PKG randomly chooses an isogeny $\varphi \in \operatorname{End}(E)$ as its master key and maps $I D_{B o b} \in\{0,1\}^{*}$ to a point $Q_{I D} \in E[n]$ by using algorithm $\mathcal{A}$. The PKG computes a public key as follows:

$$
p k_{P K G}=\varphi\left(Q_{I D}\right)
$$

and publishes $\left\{E, q, \varphi\left(Q_{I D}\right)\right\}$.
2. Extraction: The PKG computes Bob's private key $s k_{B o b}=\hat{\varphi}\left(Q_{I D}\right)$, and sends it to Bob.
3. Encryption: Alice encrypts the message $M$ using Bob's public key, $I D_{B o b}$, by performing the following steps:
a) She uses algorithm $\mathcal{A}$ to map $I D_{\text {Bob }}$ into the point $Q_{I D} \in E[n]$.
b) She chooses an isogeny $\psi \in \operatorname{End}(E)$.
c) She sets the ciphertext to be $C=(u, v)$, where

$$
u=\psi\left(Q_{I D}\right), v=e_{n}\left(\varphi\left(Q_{I D}\right), \hat{\psi}\left(Q_{I D}\right)\right)+M
$$

then she sends $C=(u, v)$ to Bob.
4. Decryption: Upon receiving $C=(u, v)$, Bob computes

$$
\begin{aligned}
e_{n}\left(u, s k_{B o b}\right) & =e_{n}\left(\psi\left(Q_{I D}\right), \hat{\varphi}\left(Q_{I D}\right)\right) \\
& =e_{n}\left(Q_{I D}, \hat{\psi}\left(\hat{\varphi} Q_{I D}\right)\right) \\
& =e_{n}\left(Q_{I D}, \hat{\varphi}\left(\hat{\psi}\left(Q_{I D}\right)\right)\right. \\
& =e_{n}\left(\varphi\left(Q_{I D}\right), \hat{\psi}\left(Q_{I D}\right)\right)
\end{aligned}
$$

and extracts the original message $M=v-e_{n}\left(\psi\left(Q_{I D}\right), \hat{\varphi}\left(Q_{I D}\right)\right)$ as required.

## 5. Security analysis

In this section, we analyze the security of our proposed scheme, which is based on the hardness of some isogeny problems as stated in the following.

Problem 1 (Isogeny Problem): For two given isogenous elliptic curves $E_{1}$ and $E_{2}$, find an isogeny $\varphi: E_{1} \rightarrow E_{2}$.

Problem 2 (Isogeny Logarithm Problem): Let $E_{1}$ and $E_{2}$ be two isogenous elliptic curves, $P \in E_{1}$ and $Q \in E_{2}$. Find an isogeny $\varphi: E_{1} \rightarrow E_{2}$ such that $Q=\varphi(P)$.

Problem 1 is a hard problem that has been studied by many researchers $[2,3,4,6$, 13]. The hardness of this problem over ordinary curves is as hard as the discrete logarithm problem, so its security is at the same level. Problem 2 is even harder than problem 1 because it must satisfy the extra term $Q=\varphi(P)$.

Generally, as mentioned earlier, there is no efficient algorithm to find an isogeny between two elliptic curves and it seems hard to determine the structure of $\operatorname{Hom}\left(E_{1}, E_{2}\right)$ and also $\operatorname{End}(E)$. Furthermore according to isogeny logarithm problem there is no efficient algorithm to find an isogeny $\varphi$ by having $P$ and $Q=\varphi(P)$.
Forward secrecy: Recall that in our proposed scheme, the public parameters are $\left\{E, \mathbb{F}_{q}, \varphi\left(Q_{I D}\right)\right\}$. Suppose Eve (the adversary) knows $p k_{P K G}=\varphi\left(Q_{I D}\right)$. To extract a message $M$, he must compute $e\left(\varphi\left(Q_{I D}\right), \hat{\psi}\left(Q_{I D}\right)\right)$. But having $E$, he could get no knowledge of isogeny $\varphi \in \operatorname{End}(E)$. Without the knowledge, this is exactly an isogeny problem that Eve is not able to solve, hence he cannot compute $e\left(\varphi\left(Q_{I D}\right), \hat{\psi}\left(Q_{I D}\right)\right)$.

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M. Bahramian<br>Department of Pure Mathematics, Faculty of Mathematical Sciences, University of Kashan, Kashan, I. R. Iran<br>e-mail: bahramianh@kashanu.ac.ir<br>E. Hajirezaei<br>Department of Pure Mathematics, Faculty of Mathematical Sciences, University of Kashan, Kashan, I. R. Iran<br>e-mail: elhamhajrezaei021@gmail.com

# NEW SUBCLASS OF MEROMORPHIC FUNCTIONS BY THE GENERALIZATION OF THE $q$-DERIVATIVE OPERATOR 

Mohammad Hassn Golmohammadi, Shahram Najafzadeh and Mohammad Reza Foroutan

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Abstract. In this paper, we introduce a new subclass of meromorphic functions, using the exponent $q$-derivative operator. Afterwards, coefficient estimates, extreme points, convex linear combination, radii of starlikeness and convexity and finally partial sum property have been investigated.
Keywords: Meromorphic functions; $q$-derivative; coefficient bound; extreme point; convex set; Hadamard product.

## 1. Introduction

Fractional calculus have started to appear more and more frequently for the modelling of relevant systems in several fields of applied sciences. For more details, one may refer to the books $[6,7,9]$ and the recent papers on the subject. The theory of $q$-analysis has attracted a considerable effort of researches due to its application in many branches of mathematics and physics and $q$-theory has an important role in various branches of mathematics and physics as for example, in the areas of special functions, ordinary fractional calculus, optimal control problems, $q$-difference, $q$-integral equations, $q$-transform analysis and in quantum physics (see for instance, $[1,2,3,4,5,8,10,16])$.
The theory of univalent functions can be described by using the theory of the $q$ calculus. Moreover, in recent years, such $q$-calculus as the $q$-integral and $q$-derivative have been used to construct several subclasses of analytic functions (see, for example, $[12,13,14,15,17])$.
Let $\Sigma$ denote the class of meromorphic functions of the form

$$
\begin{equation*}
f(z)=\frac{1}{z}+\sum_{k=1}^{+\infty} a_{k} z^{k-1} \tag{1.1}
\end{equation*}
$$

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which are analytic in the punctured unit disk

$$
\triangle^{*}=\{z \in \mathbb{C}: 0<|z|<1\}
$$

Gasper and Rahman [7] defined the $q$ - derivative of a function $f(z)$ of the form equation 1.1 by

$$
\begin{equation*}
D_{q} f(z)=\frac{f(q z)-f(z)}{(q-1) z} \tag{1.2}
\end{equation*}
$$

where $z \in \triangle^{*}$ and $0<q<1$.
Therefore, the $q$ - derivative of $f(z)=z^{k-1}$ is given by

$$
D_{q} z^{k-1}=\frac{(z q)^{k-1}-z^{k-1}}{(q-1) z}=[k-1]_{q} z^{k-2}
$$

Our aim in this paper is to introduce a new operator and a new class of functions given by equation 1.1. So we have

$$
\begin{gather*}
D_{q}^{n} f(z)=\frac{(-1)^{n} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)}{q^{\binom{n+1}{n-1}} z^{n+1}}+\sum_{k=1}^{\infty} \prod_{i=1}^{n}[k-i]_{q} a_{k} z^{k-n-1}  \tag{1.3}\\
\left(z \in \triangle^{*}, n \in \mathbb{N}=\{1,2, \cdots\}\right)
\end{gather*}
$$

where

$$
\begin{equation*}
\prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)=(1+q)\left(1+q+q^{2}\right) \cdots\left(1+q+\cdots+q^{k-1}\right) \tag{1.4}
\end{equation*}
$$

and

$$
\begin{equation*}
\prod_{i=1}^{n}[k-i]_{q}:=\left(\frac{1-q^{k-1}}{1-q}\right)\left(\frac{1-q^{k-2}}{1-q}\right) \cdots\left(\frac{1-q^{k-n}}{1-q}\right) \tag{1.5}
\end{equation*}
$$

also $\prod_{i=1}^{n}[k-i]_{q} \rightarrow \prod_{i=1}^{n}(k-i)$ as $q \rightarrow \overline{1}$. So we conclude

$$
\lim _{q \rightarrow \overline{1}} D_{q}^{n} f(z)=f^{(n)}(z) \quad, \quad z \in \triangle^{*}
$$

see also [11].
For $n \in \mathbb{N}, 0<q<1,0 \leq \lambda \leq 1,0<\alpha \leq 1$ and $\beta>0$, let $\sum_{q}(n ; \lambda, \alpha, \beta)$ be the subclass of $\sum$ consisting of functions $f$ of the form equation 1.1 and satisfying the condition

$$
\begin{align*}
& \left|\frac{z^{n+3}\left(D_{q}^{n} f(z)\right)^{\prime \prime}+z^{n+2}\left(D_{q}^{n} f(z)\right)^{\prime}-\frac{(-1)^{n}(n+1)^{2}}{q^{\binom{n+1}{n-1}}} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)}{\lambda z^{n+1}\left(D_{q}^{n} f(z)\right)+\frac{(-1)^{n} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)}{q^{\binom{n+1}{n-1}}}+\frac{(1+\lambda) \alpha}{q^{\binom{n+1}{n-1}}}}\right|  \tag{1.6}\\
& <\beta .
\end{align*}
$$

We also derive some results given various coefficient inequalities, Radii condition and Hadamard product.

## 2. Main Results

Unless otherwise mentioned, we suppose throughout this paper that $n \in \mathbb{N}$, $0<q<1,0 \leq \lambda<1,0<\alpha<1$ and $\beta>0$. First we state coefficient estimates on the class $\sum_{q}(n ; \lambda, \alpha, \beta)$.
Theorem 2.1. Let $f(z) \in \sum$, then $f(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$ is and only if

$$
\begin{align*}
& \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) a_{k}  \tag{2.1}\\
\leq & \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2} \cdots+q^{k-1}\right)-\alpha\right)}{\left.q^{(n+1} n-1\right)}
\end{align*}
$$

Proof. Let $f(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$, then equation 1.6 holds true. So by replacing equation 1.3 in equation 1.6 we have

$$
\begin{aligned}
& \left|\frac{\left.\sum_{k=1}^{+\infty}\left(\prod_{i=1}^{n}[k-i]_{q}(k-n-1)(k-n-2)+\prod_{i=1}^{n}[k-i]_{q}(k-n-1)\right) a_{k} z^{k}\right)}{\frac{(1+\lambda)}{q^{(n+1)} n-1}(-1)^{n} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)+\lambda \sum_{k=1}^{\infty} \prod_{i=1}^{n}[k-i]_{q} a_{k} z^{k}+\frac{(1+\lambda) \alpha}{q^{(n+1} n-1}}\right| \\
& <\beta \text {. }
\end{aligned}
$$

or

$$
\begin{aligned}
& \left|\frac{\sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}(k-n-1)^{2} a_{k} z^{k}}{\frac{(1+\lambda)}{\left.q^{n+1}\right)}\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)-\lambda \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q} a_{z} z^{k}}\right| \\
& <\beta .
\end{aligned}
$$

Since $\operatorname{Re}(z) \leq|z|$ for all $z$, therefore
$\left.\operatorname{Re}\left\{\frac{\sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}(k-n-1)^{2} a_{k} z^{k}}{\left.\frac{(1+\lambda)}{q^{(n+1} n-1}\right)}\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)-\lambda \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q} a_{z} z^{k}\right]\right\}$ $<\beta$.
By letting $z \rightarrow \overline{1}$ through real values, we have

$$
\begin{aligned}
& \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) a_{k} \\
& \left.\leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right)
\end{aligned}
$$

Conversely, Let equation 2.1 holds true, by equation 1.6 it is enough to show that

$$
\begin{aligned}
& X(f)= \\
& \left|\frac{z^{n+3}\left(D_{q}^{n} f(z)\right)^{\prime \prime}+z^{n+2}\left(D_{q}^{n} f(z)\right)^{\prime}-\frac{(-1)^{n}(n+1)^{2}}{q^{\binom{n+1}{n-1}}} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)}{\lambda z^{n+1}\left(D_{q}^{n} f(z)\right)+\frac{(-1)^{n} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)}{q^{\binom{n+1}{n-1}}}+\frac{(1+\lambda) \alpha}{q^{\binom{n+1}{n-1}}}}\right| \\
& <\beta,
\end{aligned}
$$

or

$$
\begin{aligned}
& \quad X(f)= \\
& \quad\left|z^{n+3}\left(D_{q}^{n} f(z)\right)^{\prime \prime}+z^{n+2}\left(D_{q}^{n} f(z)\right)^{\prime}-\frac{(-1)^{n}(n+1)^{2}}{q^{n+1} n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)\right| \\
& - \\
& -\beta \left\lvert\, \lambda z^{n+1}\left(D_{q}^{n} f(z)\right)+\frac{(-1)^{n} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)}{q^{\binom{n+1}{n-1}}}+\frac{(1+\lambda) \alpha}{\left.q^{\binom{n+1}{n-1}} \right\rvert\,}\right. \\
& < \\
& 0 .
\end{aligned}
$$

But for $0<|z|=r<1$ we have

$$
\begin{aligned}
X(f) & =\left|\sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}(k-n-1)^{2} a_{k} z^{k}\right| \\
& -\beta \left\lvert\, \frac{(1+\lambda)}{q^{(n+1} n-1}\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)\right. \\
& -\lambda \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q} a_{k} z^{k} \mid \\
& \leq \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}(k-n-1)^{2}\left|a_{k}\right| r^{k} \\
& \left.-\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right) \\
& +\lambda \beta \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left|a_{k}\right| r^{k} \leq \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}-\lambda \beta\right)\left|a_{k}\right| r^{k} \\
& -\frac{\left.\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)\right)}{\left.q^{(n+1} n-1\right)} .
\end{aligned}
$$

Since the above inequality holds for all $r(0<r<1)$, by letting $r \rightarrow \overline{1}$ and using equation 2.1 we obtain $X(f) \leq 0$, and this completes the proof.

Corollary 2.1. If function $f(z)$ of the form equation 1.1 belongs to $\sum_{q}(n ; \lambda, \alpha, \beta)$ then

$$
a_{k} \leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)} .
$$

This result is sharp for $H(z)$ given by

$$
H(z)=\frac{1}{z}+\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)} z^{k-1}
$$

Next we obtain extreme points and convex linear combination property for $f(z)$ belongs to $\sum_{q}(n ; \lambda, \alpha, \beta)$.

Theorem 2.2. The function $f(z)$ of the form equation 1.1 belongs to $\sum_{q}(n ; \lambda, \alpha, \beta)$ if and only if it can be expressed by $f(z)=\sum_{k=0}^{\infty} \sigma_{k} f_{k}(z), \sigma_{k} \geq 0, \sum_{k=0}^{\infty} \sigma_{k}=1$ where $f_{0}(z)=\frac{1}{z}$ and

$$
f_{k}(z)=\frac{1}{z}+\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) \quad z^{k-1},(k=1,2, \ldots) .
$$

Proof. Let

$$
\begin{aligned}
& f(z)=\sum_{k=0}^{\infty} \sigma_{k} f_{k}(z) \\
&=\sigma_{0} f_{0}(z) \\
&+\sum_{k=1}^{\infty} \sigma_{k}\left[\frac{1}{z}+\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right. \\
&\left.z^{k-1}\right] \\
&=\frac{1}{z}+\sum_{k=1}^{\infty} \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)} \sigma_{k} z^{k-1}
\end{aligned}
$$

Now by using Theorem 2.1 we conclude that $f(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$.
Conversely, if $f(z)$ given by equation 1.1 belongs to $\sum_{q}(n ; \lambda, \alpha, \beta)$, by letting $\sigma_{0}=1-\sum_{k=1}^{+\infty} \sigma_{k}$, where

$$
\sigma_{k}=\frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)}{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)} a_{k}, \quad(k=1,2, \ldots) .
$$

we conclude the required result.
Theorem 2.3. Let for $n=1,2, \ldots, m, f_{n}(z)=\frac{1}{z}+\sum_{k=1}^{+\infty} a_{k, n} z^{k-1}$ belongs to $\sum_{q}(n ; \lambda, \alpha, \beta)$, then $F(z)=\sum_{n=1}^{m} \sigma_{n} f_{n}(z)$ is also in the same class, where $\sum_{n=1}^{m} \sigma_{n}=1$. (Hence $\sum_{q}(n ; \lambda, \alpha, \beta)$ is a convex set.)

Proof. According to Theorem 2.1 for every $n=1,2, \ldots, m$ we have

$$
\begin{aligned}
& \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) a_{k, n} \\
& \leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{\left.q^{(n+1} n-1\right)}
\end{aligned}
$$

But

$$
\begin{aligned}
F(z) & =\sum_{n=1}^{m} \sigma_{n} f_{n}(z) \\
& =\sum_{n=1}^{m} \sigma_{n}\left(\frac{1}{z}+\sum_{k=1}^{\infty} a_{k, n} z^{k-1}\right) \\
& =\frac{1}{z} \sum_{n=1}^{m} \sigma_{n}+\sum_{k=1}^{\infty}\left(\sum_{n=1}^{m} \sigma_{n} a_{k, n}\right) z^{k-1} \\
& =\frac{1}{z}+\sum_{k=1}^{\infty}\left(\sum_{n=1}^{m} \sigma_{n} a_{k, n}\right) z^{k-1} .
\end{aligned}
$$

Since :

$$
\begin{aligned}
& \sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\left(\sum_{n=1}^{m} \sigma_{n} a_{k, n}\right) \\
& =\sum_{n=1}^{m} \sigma_{n}\left(\sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right) a_{k, n} \\
& \leq \sum_{n=1}^{m} \sigma_{n} \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}}} \\
& =\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}} \sum_{n=1}^{m} \sigma_{n}} \\
& =\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}}}
\end{aligned}
$$

then by Theorem 2.1 the proof is complete.

## 3. Radii condition and partial sum property

In this section we obtain radii of starlikeness and convexity and investigate about partial sum property.

Theorem 3.1. if the function $f(z)$ defined by equation 1.1 is in the class $\sum_{q}(n ; \lambda, \alpha, \beta)$, then $f(z)$ is meromorphically univalent starlike of order $\gamma$ in disk $|z|<R_{1}$, and it is meromorphically univalent convex of order $\gamma$ in disk $|z|<R_{2}$ where

$$
\begin{equation*}
R_{1}=\inf _{k}\left\{\frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)(1-\gamma)}{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)(k+1+\gamma)}\right\}^{\frac{1}{k}} \tag{3.1}
\end{equation*}
$$

$$
\begin{equation*}
R_{2}=\inf _{k}\left\{\frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)(1-\gamma)}{\beta(k-1)(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)(k+1+\gamma)}\right\}^{\frac{1}{k}} \tag{3.2}
\end{equation*}
$$

Proof. For starlikeness it is enough to show that

$$
\left|\frac{z f(z)^{\prime}+f(z)}{f(z)}\right|<1-\gamma
$$

but

$$
\left|\frac{z f(z)^{\prime}+f(z)}{f(z)}\right|=\left|\frac{\sum_{k=1}^{+\infty} k a_{k} z^{k}}{1+\sum_{k=1}^{+\infty} a_{k} z^{k}}\right| \leq \frac{\sum_{k=1}^{+\infty} k a_{k}|z|^{k}}{1-\sum_{k=1}^{+\infty} a_{k}|z|^{k}} \leq 1-\gamma
$$

or

$$
\sum_{k=1}^{+\infty} k a_{k}|z|^{k} \leq(1-\gamma)-(1-\gamma) \sum_{k=1}^{+\infty} a_{k}|z|^{k}
$$

or

$$
\begin{equation*}
\sum_{k=1}^{+\infty} \frac{k+1-\gamma}{1-\gamma} a_{k}|z|^{k} \leq 1 \tag{3.3}
\end{equation*}
$$

By using equation 2.1 and equation 3.3 we obtain

$$
\frac{k+1-\gamma}{1-\gamma}|z|^{k} \leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n-1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)}
$$

So, it is enough to suppose

$$
|z|^{k} \leq \frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)(1-\gamma)}{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)(k+1-\gamma)}
$$

Hence we get the required result equation 3.1. For convexity, by using the Alexander,s Theorem(If $f$ be an analytic function in the unit disk and normalized by $f(0)=f^{\prime}(0)-1=0$, then $f(z)$ is convex if and only if $z f^{\prime}(z)$ is starlike.) and applying an easy calculation we conclude the required result equation 3.2. So the proof is complete.

Theorem 3.2. Let $f(z) \in \sum$, and define

$$
\begin{equation*}
S_{1}(z)=\frac{1}{z} \quad, \quad S_{m}(z)=\frac{1}{z}+\sum_{k=1}^{m-1} a_{k} z^{k-1} \quad, \quad(m=2,3, \ldots) \tag{3.4}
\end{equation*}
$$

Also suppose $\sum_{k=1}^{+\infty} x_{k} a_{k} \leq 1$, where

$$
\begin{equation*}
x_{k}=\frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)}{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}, \tag{3.5}
\end{equation*}
$$

then

$$
\begin{equation*}
\operatorname{Re}\left(\frac{f(z)}{S_{m}(z)}\right)>1-\frac{1}{x_{m}} \quad, \quad \operatorname{Re}\left(\frac{S_{m}(z)}{f(z)}\right)>\frac{x_{m}}{1+x_{m}} \tag{3.6}
\end{equation*}
$$

Proof. Since $\sum_{k=1}^{+\infty} x_{k} a_{k} \leq 1$, they by Theorem 2.1, $f(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$.
Also by equation 1.4 and equation 1.5 we have

$$
\frac{\prod_{i=1}^{n}[k-i]_{q}}{(-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha} \geq 1
$$

so

$$
x_{k}>\frac{q^{\binom{n+1}{n-1}}\left((k-n-1)^{2}+\lambda \beta\right)}{\beta(1+\lambda)}
$$

and $\left\{x_{k}\right\}$ is an increasing sequence, therefore we obtain

$$
\begin{equation*}
\sum_{k=1}^{m-1} a_{k}+x_{m} \sum_{k=m}^{+\infty} a_{k} \leq 1 \tag{3.7}
\end{equation*}
$$

Now by putting

$$
X(z)=x_{m}\left[\frac{f(z)}{S_{m}(z)}-\left(1-\frac{1}{x_{m}}\right)\right]
$$

and making use of equation 3.7 we obtain

$$
\operatorname{Re}\left(\frac{X(z)-1}{X(z)+1}\right) \leq\left|\frac{X(z)-1}{X(z)+1}\right|=\left|\frac{x_{m} f(z)-x_{m} S_{m}(z)}{x_{m} f(z)-x_{m} S_{m}(z)+2 S_{m}(z)}\right|
$$

By a simple calculation we get $\operatorname{Re}(X(z))>0$, therefore $\operatorname{Re}\left(\frac{X(z)}{x_{m}}\right)>0$, or equivalently $\operatorname{Re}\left[\frac{f(z)}{S_{m}(z)}-\left(1-\frac{1}{x_{m}}\right)\right]>0$, and this gives the first inequality in equation 3.6. For the second inequality we consider

$$
Y(z)=\left(1+x_{m}\right)\left[\frac{S_{m}(z)}{f(z)}-\frac{x_{m}}{1+x_{m}}\right]
$$

and by using equation 3.7 we have $\left|\frac{Y(z)-1}{Y(z)+1}\right| \leq 1$, and Hence $\operatorname{Re}(Y(z))>0$, therefore $\operatorname{Re}\left(\frac{Y(z)}{1+x_{m}}\right)>0$, or equivalently $\operatorname{Re}\left[\frac{S_{m}(z)}{f(z)}-\frac{x_{m}}{1+x_{m}}\right]>0$, and this shows the second inequality in equation 3.6. So the proof is complete.

## 4. Some properties of $\sum_{q}(n ; \lambda, \alpha, \beta)$

Theorem 4.1. Let $f(z), g(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$ and given by $f(z)=\frac{1}{z}+\sum_{k=1}^{+\infty} a_{k} z^{k-1}, g(z)=\frac{1}{z}+\sum_{k=1}^{+\infty} b_{k} z^{k-1}$. Then the function
$h(z)=\frac{1}{z}+\sum_{k=1}^{+\infty}\left(a_{k}^{2}+b_{k}^{2}\right) z^{k-1}$ is also in $\sum_{q}(n ; \gamma, \alpha, \beta)$ where $\gamma \leq \frac{\lambda}{2}-\frac{(k-n-1)^{2}}{2 \beta}$.
Proof. Since $f(z), g(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$ therefore we have

$$
\left.\begin{array}{rl} 
& \sum_{k=1}^{\infty}\left[\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right]^{2} a_{k}^{2} \\
\leq & {\left[\sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) a_{k}\right]^{2}} \\
\leq & {\left[\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right)}
\end{array}\right]^{2}
$$

and

$$
\begin{aligned}
& \sum_{k=1}^{+\infty}\left[\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right]^{2} b_{k}^{2} \\
\leq & {\left[\sum_{k=1}^{+\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) b_{k}\right]^{2} } \\
& \leq\left[\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}}}\right]^{2}
\end{aligned}
$$

The above inequalities yield us

$$
\left.\begin{array}{l}
\sum_{k=1}^{\infty} \frac{1}{2}\left[\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right]^{2}\left(a_{k}^{2}+b_{k}^{2}\right) \\
\leq\left[\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right)
\end{array}\right]^{2} .
$$

Now we must show

$$
\left.\begin{array}{l}
\sum_{k=1}^{\infty}\left[\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\gamma \beta\right)\right]^{2}\left(a_{k}^{2}+b_{k}^{2}\right) \\
\leq\left[\frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right)
\end{array}\right]^{2} . . ~ \$
$$

But above inequalities holds if

$$
\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\gamma \beta\right) \leq \frac{1}{2}\left[\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right]
$$

or equivalently

$$
2(k-n-1)^{2}+2 \gamma \beta \leq(k-n-1)^{2}+\lambda \beta
$$

or

$$
\gamma \leq \frac{\lambda}{2}-\frac{(k-n-1)^{2}}{2 \beta}
$$

Theorem 4.2. The class $\sum_{q}(n ; \lambda, \alpha, \beta)$ is a convex set.

Proof. Let

$$
f(z)=\frac{1}{z}+\sum_{k=1}^{\infty} a_{k} z^{k-1}
$$

and

$$
g(z)=\frac{1}{z}+\sum_{k=1}^{\infty} b_{k} z^{k-1}
$$

be in the class $\sum_{q}(n ; \lambda, \alpha, \beta)$. For $t \in(0,1)$, it is enough to show that the function $h(z)=(1-t) f(z)+t g(z)$ is in the class $\sum_{q}(n ; \lambda, \alpha, \beta)$. Since

$$
h(z)=\frac{1}{z}+\sum_{k=1}^{\infty}\left((1-t) a_{k}+t b_{k}\right) z^{k-1}
$$

then

$$
\begin{aligned}
& \sum_{k=1}^{\infty}\left[\prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)\right]\left((1-t) a_{k}+t b_{k}\right) \\
& \left.\leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right)
\end{aligned}
$$

so $h(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$.

Corollary 4.1. Let $f_{j}(z)(j=1,2, \ldots, n)$, defined by $f_{j}(z)=\frac{1}{z}+\sum_{k=1}^{\infty} a_{k, j} z^{k-1}$ be in the class $\sum_{q}(n ; \lambda, \alpha, \beta)$, then the function $F(z)=\sum_{j=1}^{n} c_{j} f_{j}(z)$ is also in $\sum_{q}(n ; \lambda, \alpha, \beta)$ where $\sum_{j=1}^{n} c_{j}=1$.

## 5. Hadamard product

For the functions $f(z), g(z) \in \Sigma$ is given by equation 1.1, we denote by $(f * g)(z)$ the Hadamard product (or convolution) of the functions $f(z), g(z)$, that is

$$
(f * g)(z)=\frac{1}{z}+\sum_{k=1}^{+\infty} a_{k} b_{k} z^{k-1}=(g * f)(z)
$$

Theorem 5.1. If $f(z), g(z)$ defined by equation 1.1 is in the class $\sum_{q}(n ; \lambda, \alpha, \beta)$ then $(f * g)(z)=\frac{1}{z}+\sum_{k=1}^{+\infty} a_{k} b_{k} z^{k-1}$ in the class $\sum_{q}(n ; \gamma, \alpha, \beta)$ where

$$
\gamma \leq \frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)^{2}}{\beta^{2}(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}-\frac{(k-n-1)^{2}}{\beta} .
$$

Proof. Since $f(z), g(z) \in \sum_{q}(n ; \lambda, \alpha, \beta)$, so by equation 2.1
$\sum_{k=1}^{\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) a_{k} \leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}}}$
and

$$
\begin{align*}
& \sum_{k=1}^{\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) b_{k} \\
& \left.\leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{(n+1} n-1}\right) \tag{5.2}
\end{align*}
$$

By using the equation 5.1, equation 5.2 and Cauchy-Schwarts inequality we have

$$
\begin{align*}
& \sum_{k=1}^{\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) \sqrt{a_{k} b_{k}} \\
& \leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}}} \tag{5.3}
\end{align*}
$$

we must find the smallest $\gamma$ such that

$$
\begin{align*}
& \sum_{k=1}^{\infty} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\gamma \beta\right) a_{k} b_{k} \\
& \leq \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}}} \tag{5.4}
\end{align*}
$$

Now it is enough to show that

$$
\begin{align*}
& \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\gamma \beta\right) a_{k} b_{k} \\
& \leq \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right) \sqrt{a_{k} b_{k}} \tag{5.5}
\end{align*}
$$

or equivalently

$$
\sqrt{a_{k} b_{k}} \leq \frac{(k-n-1)^{2}+\lambda \beta}{(k-n-1)^{2}+\gamma \beta}
$$

But from equation 5.3,

$$
\sqrt{a_{k} b_{k}} \leq \frac{\beta(1+\lambda)\left((-1)^{n-1}\left(\prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)\right.}{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)}
$$

so it is enough that

$$
\begin{align*}
& \frac{\beta(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)}{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)} \\
& \leq \frac{(k-n-1)^{2}+\lambda \beta}{(k-n-1)^{2}+\gamma \beta} \tag{5.6}
\end{align*}
$$

By using the equation 5.6 we have

$$
\begin{align*}
\gamma & \leq \frac{q^{\binom{n+1}{n-1}} \prod_{i=1}^{n}[k-i]_{q}\left((k-n-1)^{2}+\lambda \beta\right)^{2}}{\beta^{2}(1+\lambda)\left((-1)^{n-1} \prod_{k=1}^{n}\left(1+q+q^{2}+\cdots+q^{k-1}\right)-\alpha\right)} \\
& -\frac{(k-n-1)^{2}}{\beta} \tag{5.7}
\end{align*}
$$

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Mohammad Hassn Golmohammadi
Faculty of Mathematical Sciences
Department of Pure Mathematics
Payame Noor University, P. O. Bax: 19395-3697, Tehran, Iran,
golmohamadi@pnu.ac.ir

Shahram Najafzadeh
Faculty of Mathematical Sciences
Department of Pure Mathematics
Payame Noor University, P. O. Bax: 19395-3697, Tehran, Iran, najafzadeh1234@yahoo.ie

Mohammad Reza Foroutan
Faculty of Mathematical Sciences
Department of Pure Mathematics
Payame Noor University, P. O. Bax: 19395-3697, Tehran, Iran,
foroutan ${ }_{\text {m }}$ ohammadreza@yahoo.com

# NONLINEAR NEUTRAL CAPUTO-FRACTIONAL DIFFERENCE EQUATIONS WITH APPLICATIONS TO LOTKA-VOLTERRA NEUTRAL MODEL 

Mouataz Billah Mesmouli, Abdelouaheb Ardjouni and Ahcene Djoudi

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Abstract. In this paper, we consider a nonlinear neutral fractional difference equations. By applying Krasnoselskii's fixed point theorem, sufficient conditions for the existence of solutions are established. Also, the uniqueness of a solution is given. As an application of the main theorems, we provide the existence and uniqueness of the discrete fractional Lotka-Volterra model of neutral type. Our main results extend and generalize the results that are obtained in [6].
Key words: Existence and uniqueness; fractional difference equations; Krasnoselskii fixed point theorem; contraction operator; Arzela-Ascoli's theorem; neutral discrete fractional Lotka-Volterra model.

## 1. Introduction and preliminaries

Fractional difference equations have received a special attention during the last years. Indeed, some mathematicians have recently taken the lead to develop the qualitative properties of fractional difference equations. We recall, for instance, the study made by Atici et. al. [7], [8], [9] and Abdeljawad et. al. [1], [2] (see also [4], [12], [15], [19]-[23], [25] and reference therein) who developed the transform methods, properties of initial value problems and studied applications of these equations.

Let $\mathbb{N}_{0}=\left[0, T_{1}\right] \cap \mathbb{Z}$ where $T_{1} \in[2,+\infty) \cap \mathbb{Z}$. Alzabut, Abdeljawad and Baleanu [6] discussed the existence of solutions for the difference equation

$$
\left\{\begin{array}{l}
{ }^{c} \nabla_{0}^{\alpha} x(t)=f\left(t, x(t), x\left(t-\tau_{1}\right)\right), t \in \mathbb{N}_{0}  \tag{1.1}\\
x(t)=\psi(t), t \in\left[-\tau_{1}, 0\right] \cap \mathbb{Z}
\end{array}\right.
$$

where $\tau_{1} \in \mathbb{N}, \psi:\left[-\tau_{1}, 0\right] \cap \mathbb{Z} \rightarrow \mathbb{R}, f: \mathbb{N}_{0} \times \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R}$ and ${ }^{c} \nabla_{0}^{\alpha}$ denotes the Caputo's fractional difference of order $\alpha \in(0,1)$. By employing the Krasnoselskii fixed point theorem, the authors obtained existence results.

[^7]In this paper, we are interested in the analysis of qualitative theory of the problems of the existence and uniqueness of solutions to nonlinear neutral fractional difference equations

$$
\left\{\begin{array}{l}
{ }^{c} \nabla_{0}^{\alpha} x(t)=f\left(t, x(t), x\left(t-\tau_{1}\right),{ }^{c} \nabla_{0}^{\alpha} x\left(t-\tau_{2}\right)\right), t \in \mathbb{N}_{0},  \tag{1.2}\\
x(t)=\psi(t), t \in[-\tau, 0] \cap \mathbb{Z},
\end{array}\right.
$$

where $\tau_{1}, \tau_{2} \in \mathbb{N}, \tau=\max \left(\tau_{1}, \tau_{2}\right), \psi:[-\tau, 0] \cap \mathbb{Z} \rightarrow \mathbb{R}, f: \mathbb{N}_{0} \times \mathbb{R} \times \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R}$ and ${ }^{c} \nabla_{0}^{\alpha}$ denotes the Caputo's fractional difference of order $\alpha \in(0,1)$. To prove our main results, we employ the Krasnoselskii and Banach fixed point theorems and the Arzelá-Ascoli's theorem. Moreover, we apply the main theorems to the discrete fractional Lotka-Volterra of neutral type

$$
\left\{\begin{array}{l}
{ }^{c} \nabla_{0}^{\alpha} x(t)=x(t)\left[a(t)-b(t) x\left(t-\tau_{1}\right)-c(t)^{c} \nabla_{0}^{\alpha} x\left(t-\tau_{2}\right)\right], t \in \mathbb{N}_{0}  \tag{1.3}\\
x(t)=\psi(t), t \in[-\tau, 0] \cap \mathbb{Z}
\end{array}\right.
$$

where $a, b$ and $c$ are sequences fulfill some of the conditions described below, which are medically and biologically feasible.

Now, we present some basic definitions, notations and results of fractional difference calculus [16], [17] which are used throughout this paper. For any $\alpha, t \in \mathbb{R}$, the $\alpha$ rising function is defined by

$$
\begin{equation*}
t^{\bar{\alpha}}=\frac{\Gamma(t+1)}{\Gamma(t)}, t \in \mathbb{R}-\{\ldots,-2,-1,0\}, 0^{\bar{\alpha}}=0 \tag{1.4}
\end{equation*}
$$

where $\Gamma$ is the well known Gamma function satisfying $\Gamma(\alpha+1)=\alpha \Gamma(\alpha)$.
Definition 1.1. Let $x: \mathbb{N}_{0} \rightarrow \mathbb{R}, \rho(s)=s-1, \alpha \in \mathbb{R}^{+}$and $\mu>-1$. Then

1) The nabla difference of $x$ is defined by

$$
\nabla x(t)=x(t)-x(t-1), t \in \mathbb{N}_{1}=\left[1, T_{1}\right] \cap \mathbb{Z}
$$

2) The Riemann-Liouville's sum operator of $x$ of order $\alpha>0$ is defined by

$$
\begin{equation*}
\nabla_{0}^{-\alpha} x(t)=\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}, t \in \mathbb{N}_{1} \tag{1.5}
\end{equation*}
$$

3) The Riemann-Liouville's difference operator of $x$ of order $0<\alpha<1$ is defined by

$$
\begin{equation*}
{ }^{c} \nabla_{0}^{\alpha} x(t)=\nabla_{0}^{-(1-\alpha)} \nabla x(t)=\frac{1}{\Gamma(1-\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{-\alpha}} \nabla x(s), t \in \mathbb{N}_{1} \tag{1.6}
\end{equation*}
$$

4) The power rule is defined by

$$
\begin{equation*}
\nabla_{0}^{-\alpha} t^{\bar{\mu}}=\frac{\Gamma(\mu+1-\alpha)}{\Gamma(\mu+\alpha+1)} t^{\overline{\mu+\alpha}}, t \in \mathbb{N}_{1} \tag{1.7}
\end{equation*}
$$

Let $\mathbb{N}_{-\tau}=\left[-\tau, T_{1}\right] \cap \mathbb{Z}$ where $T_{1} \in[3,+\infty) \cap \mathbb{Z}$, and $B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)$ be the Banach space of all bounded sequences with respect to the maximum norm.

Definition 1.2. A set $D$ of sequences in $B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)$ is uniformly Cauchy if for every $\epsilon>0$, there exists an integer $N^{*}$ such that $|x(t)-x(s)|<\epsilon$ whenever $t, s>N^{*}$ for any $x=\{x(n)\}$ in $D$.

The following discrete version of Arzelá-Ascoli's theorem has a crucial role in the proof of our main theorems.

Theorem 1.1. Arzelá-Ascoli's theorem A bounded, uniformly Cauchy subset D of $B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)$ is relatively compact.

The proof of the main theorem is achieved by employing the following fixed point theorem.

Theorem 1.2. Krasnoselskii's fixed point theorem [10] Let $D$ be a nonempty, closed, convex and bounded subset of a Banach space $(X,\|\cdot\|)$. Suppose that $A_{1}: D \rightarrow X$ and $A_{2}: D \rightarrow X$ are two operators such that
(i) $A_{1}$ is a contraction,
(ii) $A_{2}$ is continuous and $A_{2}(D)$ resides in a compact subset of $X$,
(iii) for any $x, y \in D, A_{1} x+A_{2} y \in D$.

Then the operator $A_{1}+A_{2}$ has a fixed point $x \in D$.

## 2. Existence and uniqueness of solutions

In this section, we give the equivalence of the problem (1.2). So, by an alternative way used in [3], [5] and [14], we turn the problem (1.2) into an equivalent equation, then, the solvability of this equivalent equation implies the existence and uniqueness of solution to the problem (1.2).

Lemma 2.1. $x$ denotes a solution of the equation (1.2) if and only if it admits the following representation

$$
\begin{equation*}
x(t)=\psi(0)+\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} z_{x}(s) \tag{2.1}
\end{equation*}
$$

where $z_{x}(t)={ }^{c} \nabla_{0}^{\alpha} x(t)$ and $x(t)=\psi(t), t \in[-\tau, 0] \cap \mathbb{Z}$.
Proof. By the same way used in [3], we get for $t \in \mathbb{N}_{-\tau}$, the initial value problem (1.2) is equivalent to the following equation
(2.2) $x(t)=\psi(0)+\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f\left(t, x(s), x\left(s-\tau_{1}\right),{ }^{c} \nabla_{0}^{\alpha} x\left(s-\tau_{2}\right)\right)$.

By the techniques used in [5] and [14], let

$$
z_{x}(t)={ }^{c} \nabla_{0}^{\alpha} x(t) \text { and } x(t)=\psi(t) \text { for } t \in[-\tau, 0] \cap \mathbb{Z}
$$

Then, the equation (2.2) is equivalent to the equation (2.1), with

$$
\begin{equation*}
z_{x}(t)=f\left(t, x(t), x\left(t-\tau_{1}\right), z_{x}\left(t-\tau_{2}\right)\right) \tag{2.3}
\end{equation*}
$$

We prove our main results under the following assumptions
(A1) For $t \in \mathbb{N}_{-\tau}$,

$$
\begin{aligned}
z_{x}(t) & =f\left(t, x(t), x\left(t-\tau_{1}\right), z_{x}\left(t-\tau_{2}\right)\right) \\
& =f_{1}(t, x(t))+f_{2}\left(t, x(t), x\left(t-\tau_{1}\right)\right)+f_{3}\left(t, x, z_{x}\left(t-\tau_{2}\right)\right)
\end{aligned}
$$

where $f_{1}, f_{2}$ and $f_{3}$ are Lipschitz functions with Lipschitz constants $L_{f_{i}}$, $i=1,2,3$, with $L_{f_{3}}<1$.
(A2) For $t \in \mathbb{N}_{-\tau}$,

$$
\begin{aligned}
\left|f_{1}(t, u(t))\right| & \leq M_{1}|u(t)| \\
\left|f_{2}(t, u(t), v(t))\right| & \leq M_{2}|u(t)||v(t)| \\
\left|f_{3}(t, u(t), v(t))\right| & \leq M_{3}|u(t)||v(t)|
\end{aligned}
$$

for any positive numbers $M_{i}, i=1,2,3$.
Define the set

$$
\begin{equation*}
D=\left\{u \in B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right),\|u\| \leq r, u(t)=\psi(t) \text { for } t \in[-\tau, 0] \cap \mathbb{Z}\right\} \tag{2.4}
\end{equation*}
$$

where $r$ satisfies

$$
\begin{equation*}
|\psi(0)|+\frac{M_{1} r+M_{2} r^{2}+M_{3} L r^{2}}{\Gamma(\alpha)} C(\alpha) \leq r \tag{2.5}
\end{equation*}
$$

and $C(\alpha)=\frac{\Gamma\left(T_{1}+\alpha\right)}{\alpha \Gamma\left(T_{1}\right)}$ is a positive constant depending on the order $\alpha$ and satisfies the inequality

$$
\begin{equation*}
L_{f_{1}} C(\alpha)<\Gamma(\alpha) \tag{2.6}
\end{equation*}
$$

Lemma 2.2. Suppose that the assumption (A1) holds. Then, for $t \in \mathbb{N}_{-\tau}, z_{x}$ satisfies the following inequality

$$
\left|z_{x}(t)-z_{y}(t)\right| \leq L\|x-y\| \text { for all } x, y \in B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)
$$

where

$$
L=\frac{L_{f_{1}}+2 L_{f_{2}}+L_{f_{3}}}{1-L_{f_{3}}}
$$

Proof. For all $x, y \in B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)$, since (A1) holds, then

$$
\begin{aligned}
& \left|z_{x}(t)-z_{y}(t)\right| \\
\leq & L_{f_{1}}|x(t)-y(t)|+L_{f_{2}}|x(t)-y(t)|+L_{f_{2}}\left|x\left(t-\tau_{1}\right)-y\left(t-\tau_{1}\right)\right| \\
& +L_{f_{3}}|x(t)-y(t)|+L_{f_{3}}\left|z_{x}\left(t-\tau_{2}\right)-z_{y}\left(t-\tau_{2}\right)\right| \\
\leq & \left(L_{f_{1}}+2 L_{f_{2}}+L_{f_{3}}\right)\|x-y\|+L_{f_{3}}\left\|z_{x}-z_{y}\right\| .
\end{aligned}
$$

Thus,

$$
\left|z_{x}(t)-z_{y}(t)\right| \leq \frac{L_{f_{1}}+2 L_{f_{2}}+L_{f_{3}}}{1-L_{f_{3}}}\|x-y\|
$$

Now, to apply Krasnoselskii's fixed point 1.2 , by Lemma 2.1 can define the operators $A_{1}$ and $A_{2}$ on $D$ by

$$
\begin{equation*}
\left(A_{1} x\right)(t)=\psi(0)+\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{1}(s, x(s)) \tag{2.7}
\end{equation*}
$$

and

$$
\begin{align*}
\left(A_{2} x\right)(t)= & \frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right) \\
& +\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right) \tag{2.8}
\end{align*}
$$

It is clear that $x$ is a solution of (1.2) if it is a fixed point of the operator $A=A_{1}+A_{2}$.
Theorem 2.1. Let conditions (A1), (A2), (2.5) and (2.6) hold. Then, the equation (1.2) has a solution in the set $D$.

Proof. From the assumptions on the set $D$, one can easily see that $D$ is a nonempty, closed, convex and bounded set.

Step 1. We prove that the $A_{1}$ defined by (2.7) is contraction. We can easily see that for $x, y \in D$

$$
\begin{aligned}
& \left|\left(A_{1} x\right)(t)-\left(A_{1} y\right)(t)\right| \\
\leq & \frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}\left|f_{1}(s, x(s))-f_{1}(s, y(s))\right| \\
\leq & L_{f_{1}} \frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}|x(s)-y(s)| \\
\leq & \frac{L_{f_{1}}}{\Gamma(\alpha)}\|x-y\| \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} .
\end{aligned}
$$

By virtue of $(1.4),(1.5),(1.7)$ and since $(t-0)^{\overline{0}}=1$, one can see that

$$
\sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}(t-0)^{\overline{0}}=\Gamma(\alpha) \nabla_{0}^{-\alpha}(t-0)^{\overline{0}}=\frac{\Gamma(t+\alpha)}{\alpha \Gamma(t)}
$$

Therefore,

$$
\left|\left(A_{1} x\right)(t)-\left(A_{1} y\right)(t)\right| \leq \frac{C(\alpha)}{\Gamma(\alpha)} L_{f_{1}}\|x-y\|, t \leq T_{1}
$$

By the assumption (2.6), we conclude that $A_{1}$ is contraction mapping on $D$.
Furthermore, we obtain for $x \in D$

$$
\begin{align*}
& \left|\left(A_{1} x\right)(t)+\left(A_{2} x\right)(t)\right| \\
\leq & \left|\psi(0)+\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{1}(s, x(s))\right| \\
& +\left|\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right)\right| \\
& +\left|\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right)\right| \\
\leq & |\psi(0)|+\frac{M_{1}\|x\|+M_{2}\|x\|^{2}+M_{3} L\|x\|^{2}}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} \\
\leq & |\psi(0)|+\frac{M_{1} r+M_{2} r^{2}+M_{3} L r^{2}}{\Gamma(\alpha)} C(\alpha) \\
\leq & r \tag{2.9}
\end{align*}
$$

which implies that $A_{1} x+A_{2} x \in D$. For $x \in D$, we also get

$$
\begin{aligned}
\left|\left(A_{2} x\right)(t)\right| \leq & \left|\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right)\right| \\
& +\left|\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right)\right| \\
\leq & \frac{M_{2}\|x\|^{2}+M_{3} L\|x\|^{2}}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} \\
\leq & \frac{M_{2} r^{2}+M_{3} L r^{2}}{\Gamma(\alpha)} C(\alpha) \\
\leq & r
\end{aligned}
$$

which implies that $A_{2}(D) \subset D$.

Step 2. We prove that $A_{2}$ is continuous. Let a sequence $x_{n}$ converge to $x$. Taking the norm of $A_{2} x_{n}-A_{2} x$, we have

$$
\begin{aligned}
& \left|\left(A_{2} x_{n}\right)(t)-\left(A_{2} x\right)(t)\right| \\
\leq & \frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}\left|f_{2}\left(s, x_{n}(s), x_{n}\left(s-\tau_{1}\right)\right)-f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right)\right| \\
& +\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}\left|f_{3}\left(s, x_{n}(s), z_{x_{n}}\left(s-\tau_{2}\right)\right)-f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right)\right| \\
= & \frac{C(\alpha)}{\Gamma(\alpha)}\left(2 L_{f_{2}}+L_{f_{3}}+L_{f_{3}} L\right)\left\|x_{n}-x\right\| .
\end{aligned}
$$

Then, we conclude that whenever $x_{n} \rightarrow x, A_{2} x_{n} \rightarrow A_{2} x$. This proves the continuity of $A_{2}$. To prove that $A_{2}(D)$ resides in a compact subset of $B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)$, i.e., $A_{2}(D)$ is a relatively compact subset. We let $t_{1} \leq t_{2} \leq T_{1}$ to get

$$
\begin{aligned}
& \left|\left(A_{2} x\right)\left(t_{2}\right)-\left(A_{2} x\right)\left(t_{1}\right)\right| \\
\leq & \left.\frac{1}{\Gamma(\alpha)} \right\rvert\, \sum_{s=1}^{t_{2}}\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}} f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right) \\
& -\sum_{s=1}^{t_{1}}\left(t_{1}-\rho(s)\right)^{\overline{\alpha-1}} f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right) \mid \\
& \left.+\frac{1}{\Gamma(\alpha)} \right\rvert\, \sum_{s=1}^{t_{2}}\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}} f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right) \\
& -\sum_{s=1}^{t_{1}}\left(t_{1}-\rho(s)\right)^{\overline{\alpha-1}} f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right) \mid \\
\leq & \frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t_{1}}\left|\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}}-\left(t_{1}-\rho(s)\right)^{\overline{\alpha-1}}\right|\left|f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right)\right| \\
& +\frac{1}{\Gamma(\alpha)} \sum_{s=t_{1}+1}^{t_{2}}\left|\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}}\right|\left|f_{2}\left(s, x(s), x\left(s-\tau_{1}\right)\right)\right| \\
& +\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t_{1}}\left|\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}}-\left(t_{1}-\rho(s)\right)^{\overline{\alpha-1}}\right|\left|f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right)\right| \\
& +\frac{1}{\Gamma(\alpha)} \sum_{s=t_{1}+1}^{t_{2}}\left|\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}}\right|\left|f_{3}\left(s, x(s), z_{x}\left(s-\tau_{2}\right)\right)\right| .
\end{aligned}
$$

By the assumption (A2) and Lemma 2.2, we obtain

$$
\begin{aligned}
& \left|\left(A_{2} x\right)\left(t_{2}\right)-\left(A_{2} x\right)\left(t_{1}\right)\right| \\
\leq & \left(M_{2} r^{2}+M_{3} L r^{2}\right)\left[\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t_{1}}\left|\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}}-\left(t_{1}-\rho(s)\right)^{\overline{\alpha-1}}\right|\right. \\
& \left.+\frac{1}{\Gamma(\alpha)} \sum_{s=t_{1}+1}^{t_{2}}\left|\left(t_{2}-\rho(s)\right)^{\overline{\alpha-1}}\right|\right] .
\end{aligned}
$$

By using (1.5), we get

$$
\begin{aligned}
& \left|\left(A_{2} x\right)\left(t_{2}\right)-\left(A_{2} x\right)\left(t_{1}\right)\right| \\
\leq & \left(M_{2} r^{2}+M_{3} L r^{2}\right)\left(\left(t_{2}\right)^{\overline{0}}-\left(t_{1}\right)^{\overline{0}}+\left(t_{2}-t_{1}\right)^{\overline{0}}\right) .
\end{aligned}
$$

From (1.7), it follows that

$$
\begin{aligned}
& \left|\left(A_{2} x\right)\left(t_{2}\right)-\left(A_{2} x\right)\left(t_{1}\right)\right| \\
\leq & \frac{\left(M_{2} r^{2}+M_{3} L r^{2}\right)}{\Gamma(\alpha+1)}\left(\nabla_{0}^{-\alpha}\left(t_{2}-0\right)^{\overline{0}}-\nabla_{0}^{-\alpha}\left(t_{1}-0\right)^{\overline{0}}+\nabla_{t_{1}}^{-\alpha}\left(t_{2}-t_{1}\right)^{\overline{0}}\right) .
\end{aligned}
$$

This implies that $A_{2}(D)$ is uniformly bounded subset of $B\left(\mathbb{N}_{-\tau}, \mathbb{R}\right)$. Thus, by virtue of the discrete Arzelá-Ascoli's theorem 1.1, we conclude that $A_{2}$ is compact.

Step 3. It remains to show that for any $x, y \in D$, we have $A_{1} x+A_{2} y \in D$. If $x, y \in D$, then we have

$$
\begin{aligned}
& \left|\left(A_{1} x\right)(t)+\left(A_{2} y\right)(t)\right| \\
\leq & \left|\psi(0)+\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{1}(s, x(s))\right| \\
& +\left|\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{2}\left(s, y(s), y\left(s-\tau_{1}\right)\right)\right| \\
& +\left|\frac{1}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} f_{3}\left(s, y(s), z_{y}\left(s-\tau_{2}\right)\right)\right| \\
\leq & |\psi(0)|+\frac{M_{1}\|x\|+M_{2}\|y\|^{2}+M_{3} L\|y\|^{2}}{\Gamma(\alpha)} \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} \\
\leq & |\psi(0)|+\frac{M_{1} r+M_{2} r^{2}+M_{3} L r^{2}}{\Gamma(\alpha)} C(\alpha) \\
\leq & r
\end{aligned}
$$

which implies that $A_{1} x+A_{2} y \in D$.
By employing the Krasnoselskii fixed point theorem, we conclude that there exists $x \in D$ such that $x=A x=A_{1} x+A_{2} x$ which is a fixed point of $A$. Hence, the equation (1.2) has at least one solution in $D$.

Remark 2.1. Note that, when $f_{3} \equiv 0$ Theorem 2.1 becomes the same Theorem 3 in [6], and this confirms the generality of the results.

It is worth noting that, the authors in [6] stated that they studied the uniqueness of solutions for the equation (1.2), but in reality they did not, because Krasnoselskii's theorem only gives us the existence of solutions, it may be only a written error. So, in this paper, we will study the uniqueness of solutions as well.

Theorem 2.2. Let conditions (A1), (A2), (2.5) and

$$
\begin{equation*}
\frac{C(\alpha)}{\Gamma(\alpha)}\left(L_{f_{1}}+2 L_{f_{2}}+L_{f_{3}}+L_{f_{3}} L\right)<1 \tag{2.10}
\end{equation*}
$$

hold. Then, the equation (1.2) has a unique solution in $D$.
Proof. Since the equation (1.2) is equivalent to (2.1), for $x \in D$ define

$$
A x=A_{1} x+A_{2} x .
$$

Step 1. We must prove that $A$ maps $D$ into itself, then by the condition (2.5) and the same way in (2.9)

$$
\begin{aligned}
|(A x)(t)| & =\left|\left(A_{1} x\right)(t)+\left(A_{2} x\right)(t)\right| \\
& \leq|\psi(0)|+\frac{M_{1} r+M_{2} r^{2}+M_{3} L r^{2}}{\Gamma(\alpha)} C(\alpha) \\
& \leq r,
\end{aligned}
$$

which implies that $A x \in D$.
Step 2. We prove that $A$ is contraction. We can see that for $x, y \in D$

$$
\begin{aligned}
& |(A x)(t)-(A y)(t)| \\
\leq & \frac{L_{f_{1}}}{\Gamma(\alpha)}\|x-y\| \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}+\frac{2 L_{f_{2}}}{\Gamma(\alpha)}\|x-y\| \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}} \\
& +\frac{\left(L_{f_{3}}+L_{f_{3}} L\right)}{\Gamma(\alpha)}\|x-y\| \sum_{s=1}^{t}(t-\rho(s))^{\overline{\alpha-1}}
\end{aligned}
$$

Therefore,

$$
|(A x)(t)-(A y)(t)| \leq \frac{C(\alpha)}{\Gamma(\alpha)}\left(L_{f_{1}}+2 L_{f_{2}}+L_{f_{3}}+L_{f_{3}} L\right)\|x-y\|, t \leq T_{1}
$$

By the assumption (2.10), we conclude that $A$ is contraction mapping on $D$.
By employing the Banach fixed point theorem, we conclude that there exists a unique $x \in D$ such that $x=A x$ which is a unique fixed point of $A$. Hence, the equation (1.2) has a unique solution in $D$.

Remark 2.2. Note that, when $f_{3} \equiv 0$ Theorem 2.2 gives the uniqueness of the solution of the equation (1.1).

Now, we can replace the assumptions (A2) and (2.5) by the following, which provide us the existence and uniqueness too.
( $\overline{\mathbf{A 2}}$ ) For $t \in \mathbb{N}_{-\tau}$, we assume that $f_{1}(t, 0)=f_{2}(t, 0,0)=f_{3}(t, 0,0) \equiv 0$ and

$$
\begin{equation*}
|\psi(0)|+\frac{\left(L_{f_{1}}+2 L_{f_{2}}+L_{f_{3}}(1+L)\right) r}{\Gamma(\alpha)} C(\alpha) \leq r \tag{2.11}
\end{equation*}
$$

Then we get the following theorems.
Theorem 2.3. Let conditions (A1), ( $\overline{\text { A2 }})$, (2.11) and (2.6) hold. Then, the equation (1.2) has a solution in the set $D$.

Proof. The proof is based on the following estimation, since $f_{1}, f_{2}$ and $f_{3}$ satisfy the assumptions (A1) and ( $\overline{\mathrm{A} 2}$ ), then

$$
\begin{aligned}
\left|f_{1}(t, x(t))\right| & =\left|f_{1}(t, x(t))-f_{1}(t, 0)\right| \\
& \leq L_{f_{1}}\|x\| \\
\left|f_{2}\left(t, x(t), x\left(t-\tau_{1}\right)\right)\right| & =\left|f_{2}\left(t, x(t), x\left(t-\tau_{1}\right)\right)-f_{2}(t, 0,0)\right| \\
& \leq 2 L_{f_{2}}\|x\|
\end{aligned}
$$

and

$$
\begin{aligned}
\left|f_{3}\left(t, x(t), z_{x}\left(t-\tau_{2}\right)\right)\right| & =\left|f_{3}\left(t, x(t), z_{x}\left(t-\tau_{2}\right)\right)-f_{3}(t, 0,0)\right| \\
& \leq L_{f_{3}}\left(\|x\|+\left\|z_{x}\right\|\right) \\
& \leq L_{f_{3}}(\|x\|+L\|x\|) \\
& =L_{f_{3}}(1+L)\|x\|
\end{aligned}
$$

The remaining steps of the proof are the same as in Theorem 2.1.
Theorem 2.4. Let conditions (A1), (̄2), (2.11) and (2.10) hold. Then, the equation (1.2) has a unique solution in $D$.

Proof. The steps of the proof is given by the same way in Theorem 2.2.
Remark 2.3. The results of this paper can be carried out for the equation

$$
\left\{\begin{array}{l}
\nabla_{0}^{\alpha} x(t)=f\left(t, x(t), x\left(t-\tau_{1}\right),{ }^{c} \nabla_{0}^{\alpha} x\left(t-\tau_{2}\right)\right), t \in \mathbb{N}_{2}=\left[2, T_{1}\right] \cap \mathbb{Z},  \tag{2.12}\\
x(t)=\psi(t), t \in[-\tau, 1] \cap \mathbb{Z},
\end{array}\right.
$$

where $\tau_{1}, \tau_{2} \in \mathbb{N}, \tau=\max \left(\tau_{1}, \tau_{2}\right), T_{1} \in[4,+\infty) \cap \mathbb{Z}, \psi:[-\tau, 1] \cap \mathbb{Z} \rightarrow \mathbb{R}, f: \mathbb{N}_{0} \times \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R}$ and $\nabla_{0}^{\alpha}$ denotes the Riemann-Liouville's fractional difference of order $\alpha \in(0,1)$. The solution of the equation (2.12) has the form

$$
x(t)=\frac{t^{\overline{\alpha-1}}}{\Gamma(\alpha)} \psi(1)+\frac{1}{\Gamma(\alpha)} \sum_{s=2}^{t}(t-\rho(s))^{\overline{\alpha-1}} z_{x}(s) .
$$

## 3. Discrete fractional Lotka-Volterra model of neutral type

Because it is very interesting to study the neutral delay population model. So, the Lotka-Volterra model has been extensively investigated by many authors see ([6], [9], [13], [11], [18], [24]) and others, through different approaches. But, all the above works studied the Lotka-Volterra model, or the neutral model with integer order. Then, there is no literature on the type of discrete neutral fractional LotkaVolterra model.

In this section, we employ Theorems 2.1 and 2.2 to prove the existence and uniqueness results for the solutions of (1.3), that represents an interspecific competition in single species with $\tau$ denotes the maturity time period.

For a bounded sequence $u$ on $\mathbb{N}_{0}$, we define $u^{+}$and $u^{-}$as follows

$$
u^{-}=\inf _{t \in \mathbb{N}_{0}} u(t) \text { and } u^{+}=\sup _{t \in \mathbb{N}_{0}} u(t)
$$

and denote

$$
\begin{aligned}
f_{1}(t, x(t)) & =a(t) x(t) \\
f_{2}\left(t, x(t), x\left(t-\tau_{1}\right)\right) & =-b(t) x(t) x\left(t-\tau_{1}\right) \\
f_{3}\left(t, x, z_{x}\left(t-\tau_{2}\right)\right) & =-c(t) x(t) z_{x}\left(t-\tau_{2}\right)
\end{aligned}
$$

where the coefficients $a, b$ and $c$ satisfy the boundedness relations

$$
a^{-} \leq a(t) \leq a^{+}, b^{-} \leq b(t) \leq b^{+}, c^{-} \leq c(t) \leq c^{+}
$$

From the conditions (A1) and ( $\overline{\mathrm{A} 2}$ ), it is easy to see that

$$
L_{f_{1}}=a^{+}, L_{f_{2}}=r b^{+}, L_{f_{3}}=r c^{+} L
$$

and

$$
M_{1}=L_{f_{1}}, M_{2}=b^{+}, M_{3}=c^{+}
$$

Theorem 3.1. Let conditions (2.5), (2.6) and

$$
L_{f_{3}}=r c^{+} L<1,
$$

hold. Then, the model (1.3) has a solution in the set D.

Theorem 3.2. Let condition (2.5), (2.10) and

$$
L_{f_{3}}=r c^{+} L<1,
$$

hold. Then, the model (1.3) has a unique solution in the set $D$.

Remark 3.1. The above theorems can be extended to $n$ species neutral competitive Lotka-Volterra system of the form

$$
\left\{\begin{array}{l}
{ }^{c} \nabla_{0}^{\alpha} x_{i}(t)=x_{i}(t)\left[a_{i}(t)-\sum_{j=1}^{n} b_{i j}(t) x\left(t-\tau_{i j}\right)-\sum_{j=1}^{n} c_{i j}(t)^{c} \nabla_{0}^{\alpha} x\left(t-\tau_{i j}\right)\right], t \in \mathbb{N}_{0},  \tag{3.1}\\
x_{i}(t)=\psi_{i}(t), t \in[-\tau, 0] \cap \mathbb{Z}
\end{array}\right.
$$

where $\tau_{i j} \in \mathbb{N}, \tau=\max _{1 \leq i, j \leq n} \tau_{i j}, \alpha \in(0,1), \psi_{i}:[-\tau, 0] \cap \mathbb{Z} \rightarrow \mathbb{R}, a^{-} \leq a_{i}(t) \leq a^{+}$, $b^{-} \leq b_{i j}(t) \leq b^{+}, c^{-} \leq c_{i j}(t) \leq c^{+}, i=1,2, \ldots, n$.

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Mouataz Billah Mesmouli
Faculty of Sciences, Department of Mathematics
University of Ha'il
Kingdom of Saudi Arabia
mesmoulimouataz@hotmail.com, m.mesmouli@uoh.edu.sa

Abdelouaheb Ardjouni
Faculty of Sciences and Technology
Department of Mathematics and Informatics
University of Souk-Ahras
P.O. Box 1553

Souk-Ahras, 41000, Algeria
abd_ardjouni@yahoo.fr

Ahcene Djoudi<br>Faculty of Sciences, Department of Mathematics<br>University of Annaba<br>P.O. Box 12<br>Annaba 23000, Algeria<br>adjoudi@yahoo.com

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