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# On weak $(\sigma, \delta)$-rigid rings over Noetherian rings 

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#### Abstract

Let R be a Noetherian integral domain which is also an algebra over $\mathbb{Q}(\mathbb{Q}$ is the field of rational numbers). Let $\sigma$ be an endomorphism of $R$ and $\delta$ a $\sigma$-derivation of $R$. We recall that a ring $R$ is a weak $(\sigma, \delta)$-rigid ring if $a(\sigma(a)+\delta(a)) \in N(R)$ if and only if $a \in N(R)$ for $a \in R(N(R)$ is the set of nilpotent elements of $R)$. With this we prove that if $R$ is a Noetherian integral domain which is also an algebra over $\mathbb{Q}, \sigma$ an automorphism of $R$ and $\delta$ a $\sigma$-derivation of $R$ such that $R$ is a weak $(\sigma, \delta)$-rigid ring, then $N(R)$ is completely semiprime.


## 1 Introduction and preliminaries

Throughout this paper $R$ will denote an associative ring with identity $1 \neq 0$, unless otherwise stated. The prime radical of a ring $R$ denoted by $P(R)$ is the intersection of all prime ideals of $R$. The set of nilpotent elements of $R$ is denoted by $N(R)$. The ring of integers, the field of rational numbers, the field

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of real numbers and the field of complex numbers are denoted by $\mathbb{Z}, \mathbb{Q}, \mathbb{R}$ and $\mathbb{C}$ respectively, unless otherwise stated.

Krempa [4] introduced $\sigma$-rigid rings and proved that if $\sigma$ is a rigid endomorphism of $R$, then it is a monomorphism preserving every minimal prime ideal and annihilator in R. Several properties of $\sigma$-rigid rings have been studied in $[2,3]$. Weak $\sigma$-rigid rings were studied by Ouyang [7]. Bhat [1] gave a necessary and sufficient condition for a commutative Noetherian ring to be weak $(\sigma, \delta)$-rigid ring.

In this article we investigate weak ( $\sigma, \delta$ )-rigid rings over Noetherian rings.
Now let $R$ be a ring, $\sigma$ an endomorphism of $R$ and $\delta$ a $\sigma$-derivation of $R$. Recall that $\delta: R \rightarrow R$ an additive map such that

$$
\delta(a b)=\delta(a) \sigma(b)+a \delta(b), \text { for all } a, b \in R
$$

is called a $\sigma$-derivation of $R$.
Example 1 Let F be a field, $\mathrm{R}=\mathrm{F}[\mathrm{x}]$ be the polynomial ring over F . Then $\sigma: \mathrm{R} \rightarrow \mathrm{R}$ defined as

$$
\sigma(\mathrm{f}(\mathrm{x}))=\mathrm{f}(-\mathrm{x}) \text { is an automorphism. }
$$

Define $\delta: R \rightarrow R$ by

$$
\delta(f(x))=f(x)-\sigma(f(x))
$$

Then $\delta$ is a $\sigma$-derivation of R .

## 1.1 $\quad \sigma$-rigid ring

Recall that in Krempa [4], an endomorphism $\sigma$ of a ring $R$ is said to be rigid if $a \sigma(a)=0$ implies that $a=0$, for all $a \in R$. A ring $R$ is said to be $\sigma$-rigid if there exists a rigid endomorphism $\sigma$ of $R$.

Example 2 Let $\mathrm{R}=\left[\begin{array}{ll}\mathrm{F} & \mathrm{F} \\ 0 & \mathrm{~F}\end{array}\right]$, where F is a field. Let $\sigma: \mathrm{R} \rightarrow \mathrm{R}$ be defined by

$$
\sigma\left(\left[\begin{array}{ll}
a & b \\
0 & c
\end{array}\right]\right)=\left[\begin{array}{ll}
a & 0 \\
0 & 0
\end{array}\right] \text { for } a, b, c \in F
$$

Then it can be seen that $\sigma$ is an endomorphism of R .
Let $0 \neq \mathrm{a} \in \mathrm{F}$. Then $\left[\begin{array}{ll}0 & \mathrm{a} \\ 0 & 0\end{array}\right] \sigma\left(\left[\begin{array}{ll}0 & a \\ 0 & 0\end{array}\right]\right)=\left[\begin{array}{ll}0 & 0 \\ 0 & 0\end{array}\right]$.
But $\left[\begin{array}{ll}0 & a \\ 0 & 0\end{array}\right] \neq\left[\begin{array}{ll}0 & 0 \\ 0 & 0\end{array}\right]$.

Hence R is not a $\sigma$-rigid ring.
We recall that $\sigma$-rigid rings are reduced rings by Hong et. al. [3]. Recall that a ring R is reduced if it has no non-zero nilpotent elements. Observe that reduced rings are abelian.

### 1.2 Weak $\sigma$-rigid ring

Note that as in Ouyang [7], a ring $R$ with an endomorphism $\sigma$ is called a weak $\sigma$-rigid ring if $a \sigma(a) \in N(R)$ if and only if $a \in N(R)$ for $a \in R$.

Example 3 (Example 2.1 of [7]) Let $\sigma$ be an endomorphism of a ring R. Let

$$
A=\left\{\left.\left[\begin{array}{ccc}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right] \right\rvert\, a, b, c, d \in R\right\}
$$

be a subring of $T_{3}(\mathbb{R})$, the ring of upper triangular matrices over R . Now $\sigma$ can be extended to an endomorphism say $\bar{\sigma}$ of $\mathcal{A}$ by $\bar{\sigma}\left(\left(a_{i j}\right)\right)=\left(\sigma\left(a_{i j}\right)\right)$. Then it can be seen that $\mathcal{A}$ is a weak $\bar{\sigma}$-rigid ring.

Example 4 Let F be a field and $\mathrm{R}=\mathrm{F}(\mathrm{x})$, the field of rational polynomials in one variable $x$ over F . Then $\mathrm{N}(\mathrm{R})=\{0\}$. Let $\sigma: \mathrm{R} \rightarrow \mathrm{R}$ be an endomorphism defined by

$$
\sigma(f(x))=f(0)
$$

Then R is not a weak $\sigma$-rigid ring. For let $\mathrm{f}(\mathrm{x})=\mathrm{xa}, \mathrm{f}(0)=0$ and $\mathrm{f}(\mathrm{x}) \sigma(\mathrm{f}(\mathrm{x}))=$ $x a .0=0 \in N(R)$. But $0 \neq f(x) \notin N(R)$.

Clearly the notion of a weak $\sigma$-rigid ring generalizes that of a $\sigma$-rigid ring. Also in [5], it has been shown that if $R$ is a weak $\sigma$-rigid ring, then $N(R)$ is completely semi-prime where $R$ is a Noetherian ring and $\sigma$ an automorphism of $R$.The converse is not true.

Example 5 [6] Let F be a field, $\mathrm{R}=\mathrm{F} \times \mathrm{F}$ and $\sigma$ an automorphism of R defined by

$$
\sigma((a, b))=(b, a), \text { for } a, b \in F
$$

Then R is a reduced ring and so $\mathrm{N}(\mathrm{R})=\{0\}$ is completely semi-prime. But R is not a weak $\sigma$-rigid ring. Since $(1,0) \sigma((1,0))=(0,0) \in N(R)$, but $(1,0) \notin N(R)$.

### 1.3 Weak $(\sigma, \delta)$-rigid rings

We generalize the above mentioned notions by involving $\sigma$ and $\delta$ together as follows:

### 1.3.1 ( $\sigma, \delta$ )-ring

Definition 1 (Definition 7 of [1]) Let R be a ring, $\sigma$ an endomorphism of R and $\delta$ a $\sigma$-derivation of R . Then R is said to be a $(\sigma, \delta)$-rigid ring if

$$
a(\sigma(a)+\delta(a)) \in P(R) \text { implies that } a \in P(R) \text { for } a \in R .
$$

Definition 2 Let R be a ring, $\sigma$ an endomorphism of R and $\delta$ a $\sigma$-derivation of R . Then R is said to be a $(\sigma, \delta)$-rigid ring if

$$
a(\sigma(a)+\delta(a))=0 \text { implies that } a=0 \text { for } a \in R .
$$

Example 6 Let $\mathrm{R}=\mathbb{C}$ and $\sigma: \mathrm{R} \rightarrow \mathrm{R}$ be defined by

$$
\sigma(\mathrm{a}+\mathfrak{i b})=\mathrm{a}-\mathfrak{i b}, \text { for all } \mathrm{a}, \mathrm{~b} \in \mathbb{R} .
$$

Then $\sigma$ is an automorphism of R .
Define $\delta$ a $\sigma$-derivation of R as

$$
\delta(z)=z-\sigma(z) \text { for } z \in R .
$$

i.e., $\delta(a+i b)=a+i b-\sigma(a+i b)=a+i b-(a-i b)=2 i b$.

Let $A=a+i b$. Then $A[\sigma(A)+\delta(A)]=0$ implies that

$$
(a+i b)[\sigma(a+i b)+\delta(a+i b)]=0
$$

i.e. $(a+\mathfrak{i b})[(a-\mathfrak{i b})+2 \mathfrak{i b}]=0$ or $(a+i b)(a+\mathfrak{i b})=0$ which implies that $\mathrm{a}=0, \mathrm{~b}=0$. Therefore, $\mathrm{A}=\mathrm{a}+\mathfrak{i b}=0$. Hence R is a $(\sigma, \delta)$-rigid ring.

### 1.3.2 Weak ( $\sigma, \delta$ )-rigid rings

Definition 3 Let R be a ring. Let $\sigma$ be an endomorphism of R and $\delta$ a $\sigma$ derivation of $R$. Then $R$ is said to be a weak $(\sigma, \delta)$-rigid ring if $a(\sigma(a)+\delta(a)) \in$ $\mathrm{N}(\mathrm{R})$ implies and is implied by $\mathrm{a} \in \mathrm{N}(\mathrm{R})$ for $\mathrm{a} \in \mathrm{R}$.

Example 7 Let $\mathrm{R}=\mathbb{Z}[\sqrt{2}]$. Then $\sigma: \mathrm{R} \rightarrow \mathrm{R}$ defined as

$$
\sigma(a+b \sqrt{2})=(a-b \sqrt{2}) \text { for } a+b \sqrt{2} \in R
$$

is an endomorphism of R .
For any $s \in R$. Define $\delta_{s}: R \rightarrow R$ by

$$
\delta_{s}(a+b \sqrt{2})=(a+b \sqrt{2}) s-s \sigma(a+b \sqrt{2}) \text { for } a+b \sqrt{2} \in R
$$

Then $\delta_{s}$ is a $\sigma$-derivation of $R$. Here $N(R)=\{0\}$.
Further,

$$
(a+b \sqrt{2})\left\{\sigma(a+b \sqrt{2})+\delta_{s}(a+b \sqrt{2})\right\} \in N(R)
$$

implies that

$$
(a+b \sqrt{2})\{(a-b \sqrt{2})+(a+b \sqrt{2}) s-s \sigma(a+b \sqrt{2})\} \in N(R)
$$

or

$$
(a+b \sqrt{2})\{a-b \sqrt{2}+a s+b s \sqrt{2}-s a+s b \sqrt{2}\} \in N(R)
$$

i.e.

$$
(a+b \sqrt{2})\{a+(2 s-1) b \sqrt{2}\} \in N(R)=\{0\}
$$

which gives $a=0, b=0$. Hence $a+b \sqrt{2}=0+0 \sqrt{2} \in N(R)$. Thus $R$ is a weak $(\sigma, \delta)$-rigid ring.

With this we prove the following:
Theorem A: Let R be Noetherian, integral domain which is also an algebra over $\mathbb{Q}$. Let $\sigma$ be an automorphism of $R$ and $\delta$ a $\sigma$-derivation of $R$. Then $R$ a weak $(\sigma, \delta)$-rigid ring implies that $N(R)$ is completely semi-prime.

The statement is proven in Theorem 1, to be found below.

## 2 Proof of the main result

We have the following before we prove the main result of this paper:
Recall that an ideal I of a ring $R$ is called completely semi-prime if $a^{2} \in I$ implies that $a \in I$ for $a \in R$.
Example 8 Let $\mathrm{R}=\left[\begin{array}{ll}\mathbb{Z} & \mathbb{Z} \\ 0 & \mathbb{Z}\end{array}\right]$. Then $\mathrm{P}_{1}=\left[\begin{array}{ll}\mathbb{Z} & \mathbb{Z} \\ 0 & 0\end{array}\right]$, $\mathrm{P}_{2}=\left[\begin{array}{ll}0 & \mathbb{Z} \\ 0 & \mathbb{Z}\end{array}\right]$, $\mathrm{P}_{3}=\left[\begin{array}{ll}0 & \mathbb{Z} \\ 0 & 0\end{array}\right]$ are prime ideals of R .

It can be easily seen that $\mathrm{P}_{1}, \mathrm{P}_{2}, \mathrm{P}_{3}$ are completely semi-prime ideals.

Proposition 1 Let R be a ring, $\sigma$ an automorphism of R and $\delta$ a $\sigma$-derivation of $R$. Then for $u \neq 0, \sigma(u)+\delta(u) \neq 0$.

Proof. Let $0 \neq u \in R$, we show that $\sigma(u)+\delta(u) \neq 0$. Let for $0 \neq u$, $\sigma(u)+\delta(u)=0$. This implies that

$$
\begin{equation*}
\delta(u)=-\sigma(u), \forall 0 \neq u \in R . \tag{1}
\end{equation*}
$$

We know that for

$$
0 \neq a, 0 \neq b \in R, \delta(a b)=\delta(a) \sigma(b)+a \delta(b) .
$$

By using equation 1, we have

$$
\begin{array}{rlrl} 
& & -\sigma(a b) & =-\sigma(a) \sigma(b)+a(-\sigma(b)) \\
\Rightarrow & -\sigma(a b) & =-[\sigma(a)+a] \sigma(b) \\
\Rightarrow & \sigma(a) \sigma(b) & =-[\sigma(a)+a] \sigma(b) \\
\Rightarrow & \sigma(a) & =\sigma(a)+a
\end{array}
$$

Therefore, $a=0$ which is not possible. Hence $\sigma(u)+\delta(u) \neq 0$.
We now prove the main result of this paper in the form of the following Theorem:

Theorem 1 Let R be a Noetherian integral domain which is also an algebra over $\mathbb{Q}$. Let $\sigma$ be an automorphism of R and $\delta$ a $\sigma$-derivation of R . Then $\mathrm{R} a$ weak $(\sigma, \delta)$-rigid ring implies that $\mathrm{N}(\mathrm{R})$ is completely semi-prime.

Proof.Let $R$ be a weak ( $\sigma, \delta$ )-rigid ring. Then we will show that $N(R)$ is completely semi-prime. Let $a \in R$. Since $R$ is a weak $(\sigma, \delta)$-rigid ring. Let

$$
\{a(\sigma(a)+\delta(a))\}^{2} \in N(R) .
$$

Then there exists a positive integer $n$ such that $\left[a^{2}(\sigma(a)+\delta(a))^{2}\right]^{n}=0$ which implies that $a^{2 n}(\sigma(a)+\delta(a))^{2 n}=0$. But by Proposition (1), $\sigma(a)+\delta(a) \neq 0$. Hence $a^{2 n}=0$ which implies that $a^{n}=0$, because $R$ is an integral domain. Therefore, $a^{n}(\sigma(a)+\delta(a))^{n}=0$ or $\{a(\sigma(a)+\delta(a))\}^{n}=0$. Thus $a(\sigma(a)+$ $\delta(a)) \in N(R)$. Hence $N(R)$ is completely semi-prime.

The converse is not true.
Example 9 Let F be a field. Let $\mathrm{R}=\mathrm{F} \times \mathrm{F}$ and $\sigma$ an automorphism of R defined by

$$
\sigma((a, b))=(b, a) \text { for } a, b \in F
$$

Then R is a reduced ring and so $\mathrm{N}(\mathrm{R})=\{0\}$ and therefore, it is completely semi-prime. Let $\mathrm{r} \in \mathrm{F}$. Define $\delta_{\mathrm{r}}: \mathrm{R} \rightarrow \mathrm{R}$ by

$$
\delta_{r}((a, b))=(a, b) r-r \sigma((a, b)) \text { for } a, b \in F
$$

Then $\delta_{\mathrm{r}}$ is a $\sigma$-derivation of R . Also R is not a weak $(\sigma, \delta)$-rigid ring. For take $(1,-1) \in R, r=\frac{1}{2}$. Then

$$
\begin{aligned}
& \delta_{r}((1,-1))=(1,-1) \frac{1}{2}-\frac{1}{2} \sigma((1,-1)) \\
= & (1,-1) \text { and }(1,-1)\left[\sigma(1,-1)+\delta_{r}(1,-1)\right] \\
= & (1,-1)[(-1,1)+(1,-1)] \\
= & (1,-1)(0,0)=(0,0) \in N(R) .
\end{aligned}
$$

$$
\text { But }(1,-1) \notin \mathrm{N}(\mathrm{R}) .
$$

Corollary 1 Let R be a commutative Noetherian, integral domain which is also an algebra over $\mathbb{Q}$. Let $\sigma$ be an automorphism of R and $\delta$ a $\sigma$-derivation of $R$. Then $R$ a weak $(\sigma, \delta)$-rigid ring implies that $N(R)$ is completely semiprime.

Also we note that if $R$ is a ( $\sigma, \delta$ )-rigid ring then it is a weak ( $\sigma, \delta$ )-rigid ring, but the converse need not be true as in the following example:

Example 10 Let $\sigma$ be an endomorphism of a ring R and $\delta$ a $\sigma$-derivation of R. Let R be a $(\sigma, \delta)$-rigid ring. Then

$$
R_{3}=\left\{\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]: a, b, c, d \in R\right\}
$$

is a subring of $\mathrm{T}_{3}(\mathrm{R})$. The endomorphism $\sigma$ of R can be extended to the endomorphism $\bar{\sigma}: R_{3} \rightarrow R_{3}$ defined by $\bar{\sigma}\left(\left(a_{i j}\right)\right)=\left(\sigma\left(a_{i j}\right)\right)$ and $\delta$ can be extended to $\bar{\delta}: R_{3} \rightarrow R_{3}$ by $\bar{\delta}\left(\left(a_{i j}\right)\right)=\left(\delta\left(a_{i j}\right)\right)$. Let

$$
\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\left\{\bar{\sigma}\left(\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\right)+\bar{\delta}\left(\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\right)\right\} \in N(R) .
$$

Then there is some positive integer n such that

$$
\left(\left[\begin{array}{ccc}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\left\{\left[\begin{array}{ccc}
\sigma(a) & \sigma(b) & \sigma(c) \\
0 & \sigma(a) & \sigma(d) \\
0 & 0 & \sigma(a)
\end{array}\right]+\left[\begin{array}{ccc}
\delta(a) & \delta(b) & \delta(c) \\
0 & \delta(a) & \delta(d) \\
0 & 0 & \delta(a)
\end{array}\right]\right\}\right)^{n}=0
$$

which implies that

$$
\left(\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\left\{\left[\begin{array}{ccc}
\sigma(a)+\delta(a) & \sigma(b)+\delta(b) & \sigma(c)+\delta(c) \\
0 & \sigma(a)+\delta(a) & \sigma(d)+\delta(d) \\
0 & 0 & \sigma(a)+\delta(a)
\end{array}\right]\right\}\right)^{n}=0
$$

or
$\left[\begin{array}{ccc}a(\sigma(a)+\delta(a)) & a(\sigma(b)+\delta(b))+b(\sigma(a)+\delta(a)) & a(\sigma(c)+\delta(c))+b(\sigma(d)+\delta(d))+c(\sigma(a)+\delta(a)) \\ 0 & a(\sigma(a)+\delta(a)) & a(\sigma(d)+\delta(d))+d(\sigma(a)+\delta(a)) \\ 0 & 0 & a(\sigma(a)+\delta(a))\end{array}\right]^{n}$
$=0$, which gives

$$
a(\sigma(a)+\delta(a)) \in N(R)
$$

Since R is reduced, we have

$$
a(\sigma(a)+\delta(a))=0
$$

which implies that $\mathrm{a}=0$, since R is a $(\sigma, \delta)$-rigid ring. Hence

$$
\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]=\left[\begin{array}{lll}
0 & b & c \\
0 & 0 & d \\
0 & 0 & 0
\end{array}\right] \in N(R)
$$

Conversely, assume that

$$
\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right] \in N(R)
$$

Then there is some positive integer n such that

$$
\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]^{n}=\left[\begin{array}{ccc}
a^{n} & * & * \\
0 & a^{n} & * \\
0 & 0 & a^{n}
\end{array}\right]=0
$$

which implies that $\mathrm{a}=0$, because R is reduced (Here $*$ are non-zero terms involving summation of powers of some or all of $a, b, c, d)$. So

$$
\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\left\{\bar{\sigma}\left(\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\right)+\bar{\delta}\left(\left[\begin{array}{lll}
a & b & c \\
0 & a & d \\
0 & 0 & a
\end{array}\right]\right)\right\}
$$

$$
\begin{aligned}
& =\left[\begin{array}{lll}
0 & b & c \\
0 & 0 & d \\
0 & 0 & 0
\end{array}\right]\left\{\left[\begin{array}{ccc}
0 & \sigma(b) & \sigma(c) \\
0 & 0 & \sigma(d) \\
0 & 0 & 0
\end{array}\right]+\left[\begin{array}{ccc}
0 & \delta(b) & \delta(c) \\
0 & 0 & \delta(d) \\
0 & 0 & 0
\end{array}\right]\right\} \\
& =\left[\begin{array}{lll}
0 & b & c \\
0 & 0 & d \\
0 & 0 & 0
\end{array}\right]\left\{\left[\begin{array}{ccc}
0 & \sigma(b)+\delta(b) & \sigma(c)+\delta(c) \\
0 & 0 & \sigma(d)+\delta(d) \\
0 & 0 & 0
\end{array}\right]\right\} \\
& =\left[\begin{array}{lll}
0 & 0 & b(\sigma(d)+\delta(d)) \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right] \in N(R) .
\end{aligned}
$$

Therefore, $\mathrm{R}_{3}$ is a weak $(\bar{\sigma}, \bar{\delta})$-rigid ring. Also since $\mathrm{R}_{3}$ is not reduced, $\mathrm{R}_{3}$ is not a $(\bar{\sigma}, \bar{\delta})$-rigid ring.

## References

[1] M. Abrol, V. K. Bhat, Ore Extensions Over $(\sigma, \delta)$-Rings, Eur. J. Pure Appl. Math., 8 (4) (2015), 462-468.
[2] Y. Hirano, On the uniqueness of rings of coefficients in skew polynomial rings, Publ. Math. Debrecen, 54 (3, 4) (1999), 489-495.
[3] C. Y. Hong, N. K. Kim and T. K. Kwak, Ore extensions of Baer and p.p - rings, J. Pure Appl. Algebra, 151 (3) (2000), 215-226.
[4] J. Krempa, Some examples of reduced rings, Algebra Colloq., 3 (4) (1996), 289-300.
[5] N. Kumari, S. Gosani, V. K. Bhat, Skew polynomial rings over weak $\sigma$ rigid rings and $\sigma(*)$-rings, Eur. J. Pure Appl. Math., 6 (1) (2003), 59-65.
[6] T. K. Kwak, Prime radicals of Skew polynomial rings, Int. J. Math. Sci., 2 (2) (2003), 219-227.
[7] L. Ouyang, Extensions of generalized $\alpha$-rigid rings, Wiley, Int. Electron. J. Algebra, 3 (2008), 103-116.

# Assessing the effect size of users' consciousness for computer networks vulnerability 

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#### Abstract

In this paper the conditions and the findings of a simulation study is presented for assessing the effect size of users' consciousness to the computer network vulnerability in risky cyber attack situations at a certain business. First a simple model is set up to classify the groups of users according to their skills and awareness then probabilities are assigned to each class describing the likelihood of committing dangerous reactions in case of a cyber attack. To quantify the level of network vulnerability a metric developed in a former work is used. This metric shows the approximate probability of an infection at a given business with well specified parameters according to its location, the type of the attack, the protections used at the business etc. The findings mirror back the expected tendencies namely if the number of conscious user is on the


[^0]rise the "relative improvement of the cyber security" is increasing. The tendencies in the change of this relative improvement are established, different graphs and curves are constructed to give an overall view for the influence of the different parameters. In addition to these general conclusions assessments are made for the magnitude and for the range of the relative cyber security improvement. An interesting findings that even in the case of small differences in skills making the users more conscious in their reactions can significantly enhance the level of cyber security at a business.

## 1 Introduction

Assessing the extent of vulnerability of a net of computers against outer cyber threats is of prime interest for both the IT experts sector and the businesses using computer networks. Most of the IT solutions concentrate on different hardware and software protections against the threats and little attention is paid on the effect of the users' behavior during their daily routine handling potentially risky situations. Opening potentially dangerous websites, clicking on links in emails from unknown source, downloading files to the computers are typical "user tricks" which may raise the level of risk of infections.

Making the users more conscious in their computer usage is an evidential tool for enhancing the cyber security of a business. However to give some measure for the effectiveness of these kind of efforts (trainings for employees, incentives, penalties, etc.) is essential for those offering these services and for the managements of the businesses as well.

In this paper the findings of computer simulation studies are presented where the users at a business are categorized according to their everyday computer usage. The three categories (Naive, Typical, and Conscious) encompasses the different types of user groups mainly reflecting their attitudes and behaviors in risky cyber threat situations.

To make the influence of the different groups sensible the $p_{s}$ metric for measuring cyber vulnerability developed in [4], [3], [2] and [1] is used. This $\mathrm{p}_{\mathrm{s}}$ metric is the probability that at least one cyber malware (virus) can successfully go through the IT protections and infects a certain net of computers.

The features of the given net and the prevailing threats at the location investigated are precisely characterized by different matrices describing the presence or absence of the different types of protections, the relative frequencies and the level of danger of different viruses and the level of danger of the different user tricks (Simple, Moderate, Complex).

This $p_{s}$ metric holds the intrinsic characteristic of being monotonously increasing as the number of viruses, devices or users increases. In this study to ensure getting such a $p_{s}$ metric which reflects the real differences of vulnerability attributable exclusively to the different categories of users' consciousness (awareness) and excludes the effect of the business size the number of users, devices, and threats are kept constant while the proportions of the different types of users are changed.

Besides these kept-constant parameters other input values like probabilities of occurrences of certain threats or probabilities of applying certain user tricks, etc. were also temporarily fixed in the analysis (at the values which are typical for the majority of businesses) but later the sensitivity of the $p_{s}$ value against these varying input values is demonstrated.

The final goal was to illustrate the extent of changes in the value of $p_{s}$ when the level of consciousness at a business increases due to some measures introduced by the management. The increase of consciousness is embodied in the increase of the proportion of Conscious users reevaluating some users' status from Typical or Naive to Conscious or in the increase of the proportion of Typical users reevaluating some users' status from Naive to Typical.
The direct relationship between the $p_{s}$ value and some specific business financial indicator is not investigated here.

## 2 The model for users classification

According to their consciousness there are three distinguished class of users

- Naive users,
- Typical users,
- Conscious users.

As the names of the categories suggest Naive users are assumed to commit dangerous actions even in cases requiring very simple user tricks while Conscious users are victims only of threats requiring more sophisticated user tricks.

Let $v$ be

$$
\left[\begin{array}{l}
c \\
t \\
n
\end{array}\right]
$$

the vector where $\mathrm{c}(\mathrm{t}, \mathrm{n}$ reps.) is the number of the Conscious (Typical, Naive resp.) users. Let $r$ be the number of users. Observe $c+t+n=r$.

There are three distinguished type of user tricks

- Simple user trick,
- Moderate user trick,
- Complex user trick.

The Simple (Complex resp.) user trick is the easiest (complicated resp.) trick for the threat to attack a given device.

Let $P_{\text {skills }}$ be the

|  | Simple | Moderate | Complex |
| :--- | :---: | :---: | :---: |
| Conscious | $\beta_{\mathrm{C}, 1}$ | $\beta_{\mathrm{C}, 2}$ | $\beta_{\mathrm{C}, 3}$ |
| Typical | $\beta_{\mathrm{T}, 1}$ | $\beta_{\mathrm{T}, 2}$ | $\beta_{\mathrm{T}, 3}$ |
| Naive | $\beta_{\mathrm{N}, 1}$ | $\beta_{\mathrm{N}, 2}$ | $\beta_{\mathrm{N}, 3}$ |

$3 \times 3$ matrix. The real number $\beta_{\mathrm{C}, 1}\left(\beta_{\mathrm{C}, 2}, \beta_{\mathrm{C}, 3}\right.$ resp. $)$ is the probability that a Conscious user uses a Simple (Moderate, Complex resp.) user trick. The real number $\beta_{\mathrm{T}, 1}\left(\beta_{\mathrm{T}, 2}, \beta_{\mathrm{T}, 3}\right.$ resp. $)$ is the probability that a Typical user uses a Simple (Moderate, Complex resp.) user trick. The real number $\beta_{N, 1}\left(\beta_{N, 2}, \beta_{N, 3}\right.$ resp.) is the probability that a Naive user uses a Simple (Moderate, Complex resp.) user trick. It is assumed that $\beta_{i, 1}<\beta_{i, 2}<\beta_{i, 3}$ and $\beta_{C, j}<\beta_{T, j}<\beta_{N, j}$ for $i=C, T, N$ and $j=1,2,3$.

Combining $v$ and $\mathrm{P}_{\text {skills }}$ let $\mathrm{P}_{\text {user-usertrick }}$ be the

|  | Simple | Moderate | Comlex |
| :---: | :---: | :---: | :---: |
| $u_{1}$ | $\beta_{\mathrm{C}, 1}$ | $\beta_{\mathrm{C}, 2}$ | $\beta_{\mathrm{C}, 3}$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |
| $u_{\mathrm{c}}$ | $\beta_{\mathrm{C}, 1}$ | $\beta_{\mathrm{C}, 2}$ | $\beta_{\mathrm{C}, 3}$ |
| $\boldsymbol{u}_{\mathrm{c}+1}$ | $\beta_{\mathrm{T}, 1}$ | $\beta_{\mathrm{T}, 2}$ | $\beta_{\mathrm{T}, 3}$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |
| $u_{\mathrm{c}+\mathrm{t}}$ | $\beta_{\mathrm{T}, 1}$ | $\beta_{\mathrm{T}, 2}$ | $\beta_{\mathrm{T}, 3}$ |
| $\boldsymbol{u}_{\mathrm{c}+\mathrm{t}+1}$ | $\beta_{\mathrm{N}, 1}$ | $\beta_{\mathrm{N}, 2}$ | $\beta_{\mathrm{N}, 3}$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |
| $u_{\mathrm{c}+\mathrm{t}+\mathrm{n}}$ | $\beta_{\mathrm{N}, 1}$ | $\beta_{\mathrm{N}, 2}$ | $\beta_{\mathrm{N}, 3}$ |

$r \times 3$ matrix. Here $u_{1}, \ldots, u_{c}$ denote the users belonging to the group of Conscious users, $\mathfrak{u}_{\mathfrak{c}+1}, \ldots, \mathfrak{u}_{\mathfrak{c}+\boldsymbol{t}}$ denote the users in the Typical group and $u_{c+t+1}, \ldots, u_{c+t+n}$ stand for the users in the Naive group.

In this paper this $\mathrm{P}_{\text {user-usertrick }}$ matrix is the main tool to study the effect of users' behaviour at a business against different cyber threats. Changing the values of $c, t, n$ within the fixed value of $r(r=c+t+n)$ or changing the proportion (distribution) of probabilities in the rows/columns of this matrix enables us to study different situations and give an overview about the magnitude of the effect size of users' consciousness to the cyber vulnerability.

## 3 The $p_{s}$ probability

In [2] the probability of infection $p_{s}$ was introduced which is the probability that the investigated landscape will be infected by at least one malware. This can be calculated in the following form

$$
\begin{equation*}
p_{s}=1-\prod_{t=1, \ldots, k ; u=1, \ldots, r ; d=1, \ldots, m}\left(1-p_{\text {user }}(t, u) \cdot p_{\text {device }}(t, d) \cdot p_{\text {prev }}(t)\right) \tag{1}
\end{equation*}
$$

for any $u \in U, t \in T$ and $d \in D$, where $U$ ( $T$ and $D$ resp.) symbolizes the set of users (threats and devices resp.) at the landscape investigated. In (1) $p_{u s e r}(t, u)$ is the probability that the threat $t$ infects the landscape using at least one usertrick through the user $u$. In (1) $p_{\text {device }}(t, d)$ is the probability of a successful attempts of the threat $t$ through any protection protecting the device $d$. In (1) $p_{\text {prev }}(t)$ is the probability that an attack is in the form of the threat t .

In [2] and [1] it was shown how these probabilities can be computed using several parameters describing the present state of the investigated landscape, the prevailing threats, the devices, the state of protections etc.

Since the number of these parameters influencing the value of $p_{s}$ is numerous it is not easy to see general tendencies due to some selected specific parameters without keeping the others at constant values.

In this study where the aim was to get information about the effect of users' consciousness the parameter values referring to the other features of the investigated business have been fixed. Hence two types of parameters have been distinguished:

- kept-constant parameters and
- study parameters.

The kept-constant parameters are those which are not varied in the study at all. They can simply be regarded as those specifying the basic unit of
comparison. For example the total number of users is fixed to 100 since this parameter can be regarded as the size of the users' sample taken from the population of all users at a given business.

The study parameters are being varied in the study. Since they are still rather numerous it was practical to temporarily select and fix them at some typical real world values (called "typical study values") and investigate the sensitivity of $p_{s}$ to the deviation from these typical study values later.

## 4 The kept-constant parameters and their values in the study

The values of the kept-constant parameters:

- the number of malwares: $k=10$,
- the number of users: $\mathrm{r}=100$,
- the number of devices: $\mathfrak{m}=10$,
- the number of protections: $\mathfrak{n}=10$,
- the number of user tricks: $\mathfrak{i}=3$,
- the number of groups of user skills: $s=3$,
- the probability that an attacker will use a particular threat or class of threats against the enterprise: $\mathrm{P}_{\text {prev }}=[1 / k, 1 / k, \ldots, 1 / k]$,
- the $Z_{\text {device-elements }}$ matrix which describes that in this situation each virus can work on each device:

$$
Z_{\text {device-elements }}=\begin{array}{c|ccc} 
& 1 & \ldots & k \\
\hline 1 & 1 & \ldots & 1 \\
\vdots & \vdots & \ldots & \vdots \\
m & 1 & \ldots & 1
\end{array},
$$

- the $Z_{\text {device-prot-install }}$ matrix which describes that in this situation each protection is installed on each device:

$$
\mathrm{Z}_{\text {device-prot-install }}=\begin{array}{c|ccc} 
& 1 & \ldots & n \\
\hline 1 & 0 & \ldots & 0 \\
\vdots & \vdots & \ldots & \vdots \\
\mathrm{~m} & 0 & \ldots & 0
\end{array},
$$

- the $P_{\text {prot }}$ matrix which describes that in this situation the probability of a successful attempt of a given threat through at a given protection:

$$
P_{\text {prot }}=\begin{array}{c|ccc} 
& 1 & \ldots & n \\
\hline 1 & 1 /(n k) & \ldots & 1 /(n k) \\
\vdots & \vdots & \ldots & \vdots \\
k & 1 /(n k) & \ldots & 1 /(n k)
\end{array} .
$$

## 5 The study parameters and their "typical study values"

The "typical study values":

- The distribution of probabilities in the $\mathrm{P}_{\text {skills }}$ matrix:

$\mathrm{P}_{\text {skills }}=$|  | Simple | Moderate | Complex |
| :--- | :---: | :---: | :---: |
| Conscious | $p_{0}$ | $3 p_{0}$ | $6 p_{0}$ |
| Typical | $3 p_{0}$ | $9 p_{0}$ | $18 p_{0}$ |
| Naive | $6 p_{0}$ | $18 p_{0}$ | $36 p_{0}$ |, where $p_{0}=0.0001$.

- The distribution of probabilities in the $P_{\text {usertrick }}$ matrix:

$\mathrm{P}_{\text {usertrick }}=$|  | Simple | Moderate | Complex |
| :---: | :---: | :---: | :---: |
| $\mathrm{t}_{1}$ | $\alpha_{1}$ | $\alpha_{2}$ | $\alpha_{3}$ |
| $\vdots$ | $\vdots$ | $\vdots$ | $\vdots$ |, where |  |
| :---: |
| $\alpha_{1}=0.6$, |
| $\alpha_{2}=0.3$, |
| $t_{k}$ |
| $\alpha_{1}$ |

Here $P_{\text {usertrick }}$ is a $k \times 3$ matrix. The real number $\alpha_{1}\left(\alpha_{2}, \alpha_{3}\right.$ resp. $)$ is the probability that a threat uses a Simple (Moderate, Complex resp.) user trick. The integer number $k$ denotes the total number of threats involved in the study.

The fact that each row of the $P_{\text {usertrick }}$ matrix has the same values of $\alpha_{1}, \alpha_{2}$ and $\alpha_{3}$ tacitly assumes that each virus behaves similarly that is all viruses involved in the analysis belong to the same group of viruses with respect to their user tricks required to activate them. Obviously it does not hold for all groups of viruses so later the effect of differently distributed $\alpha$ probabilities will be investigated.

## 6 The effect of users' consciousness for the $p_{s}$ value in the case of "typical study values"

To illustrate the magnitude of the effect size of users' consciousness extreme situations have been analyzed where first all users were assumed to belong to one specific class of consciousness ("Original class") and gradually each user is trained to step up into a "higher" class of consciousness ("Improved class"). Three extreme versions of this "class change" are detailed:

- from Naive to Typical,
- from Naive to Conscious,
- from Typical to Conscious.

Various graphs have been constructed to visualize the effect size.
On the first kind of graphs the change of the absolute value of $p_{s}$ probability is depicted as the function of the number of "reevaluated" users. This probability is denoted by $p(x)$ where $x$ refers to the number of "reevaluated" users. (Sometimes they are called to "reeducated" users.) Accordingly $r-x$ refers to the number of users still in the "Original class". If $r=100, p(0)$ refers to the case when all users are in their "Original class" while $p(100)$ indicates the situation when all users have been "reevaluated".

On the second kind of graphs the $\Delta(x)$ function, the relative change of the $p_{s}$ probability is illustrated

$$
\Delta(x)=\frac{p(0)-p(x)}{p(0)}
$$

Since the absolute value of $p_{s}$ is very much dependent on the different "class change" situations more practical to calculate the ratio of the probability change to the $p_{s}$ value in the "Original class". This change can be interpreted as the "relative improvement of the defence".

From business point of view this relative improvement can be the basis of any management measures for the sake of improving cyber security through the enhancement of users' consciousness.

In Fig. 1(a) the $\mathfrak{p}(x)$ curves are shown for the three extreme "class change" situations. The changes of $\mathfrak{p}(x)$ are almost linear in all three cases, naturally the slope of the Naive $\mapsto$ Consious line is the steepest one.

On the Naive $\mapsto$ Consious curve in Fig. 1(a) it can be seen that the initial probability of the infection is $p(0)=0.01252$ and if 50 Naive users
are reeducated to Conscious users, then the probability of the infection is $p(50)=0.00732$.


Figure 1: The "typical study values".

The $\Delta(x)$ curves can be constructed from the corresponding $p(x)$ curves hence for all three "class change" situations these curves also show almost perfect linearity. If 50 Naive users are reeducated to Conscious users, then the change of the defence is $p(0)-p(50)=0.01252-0.00732=0.0052$ and the normalized change of defence with respect to the initial probability of the infection is $\Delta(50)=(p(0)-p(50)) / p(0)=0.41511$ which can be seen in Fig. 1(b).

These $\Delta(x)$ curves can be regarded as the most important findings of the simulation studies. Even for those who are not very familiar with the issues of cyber security the magnitude of the improvement can be convincing. Seeing the different extreme situations of "class change" one can find that significant improvement can be reached through the enhancement of the users' consciousness.

The range of this relative improvement for a specific $x$ value can be assessed as the difference of the $\Delta(x)$ values for the Naive $\mapsto$ Conscious and the Naive $\mapsto$ Typical "class change" situations. Both the magnitude and the range of the improvement is monotonously increasing as the number of the "reevaluated" users is increasing. For example the relative improvement of the defence can vary from about 25 to 40 percent when half of the users' status has been changed. In Fig. 1(b) this range is indicated with a thick solid line.

## 7 The sensitivity of the relative cyber security improvement to the deviations from the "typical study values"

In the remaining sections the sensitivity of the relative security improvement is investigated. As it was stated earlier the simulation studies were elaborated for those "typical study values" of the study parameters which are believed to be characteristic for real world average size businesses in every day cyber risk situations.

However it is worth to check whether slight or moderate deviations from these study values results in basically different conclusions or the findings are rather insensitive to these deviations.

### 7.1 Varying $p_{0}$

In the Fig. 2(a) - Fig. 4(b) the influence of the deviation from the $p_{0}=0.0001$ study value is shown. The range of the $p_{0}$ values goes from 0.00001 to 0.01 .


Figure 2: The value $\mathrm{p}_{0}=0.00001$.
The magnitude of the $p(x)$ probabilities varies and the shape of the $p(x)$ and $\Delta(x)$ curves are slightly different from the almost linear "typical study values" curves.

With the increase of the $p_{0}$ probability the $\Delta(x)$ curves deviates from the linear relationships and tend to be more "exponential-like". At the same time the $\Delta(x)$ values for a specific $x$ value are typically less then those for the typical $p_{0}=0.0001$ study value .


Figure 3: The value $p_{0}=0.001$.


Figure 4: The value $p_{0}=0.01$.

Besides these slight differences one can see that the range of the relative cyber security improvement still remain similar to the previously investigated cases. For example for the $p_{0}=0.01$ value the range of $\Delta(x)$ is somewhere between 15 and 30 percent if half of the users' status has been "reevaluated". In Fig. $4(\mathrm{~b})$ this range is indicated with a thick solid line.

### 7.2 Varying $\alpha_{1}, \alpha_{2}$ and $\alpha_{3}$

In the Fig. 5(a) - Fig. 5(b) the influence of the deviation from the $\alpha_{1}=0.6$, $\alpha_{2}=0.3, \alpha_{3}=0.1$ study values is shown.

The results are presented for the following sets of deviated $\alpha$ values:

- $\alpha_{1}=0, \alpha_{2}=0, \alpha_{3}=1$,
- $\alpha_{1}=0, \alpha_{2}=1, \alpha_{3}=0$,
- $\alpha_{1}=1, \alpha_{2}=0, \alpha_{3}=0$,
- $\alpha_{1}=1 / 3, \alpha_{2}=1 / 3, \alpha_{3}=1 / 3$.

These sets of $\alpha$ values can be regarded as the representations of different virus groups requiring different user tricks to activate them.


Figure 5: Varying $\alpha_{1}, \alpha_{2}$ and $\alpha_{3}$.

The $\Delta(x)$ curves practically coincide with each other convincingly demonstrating the insensitivity of the $\Delta(x)$ function to the deviations from the typical $\alpha_{1}, \alpha_{2}, \alpha_{3}$ study values.

### 7.3 Varying the ratio of the columns of $\mathrm{P}_{\text {skills }}$

To check the sensitivity of the relative improvement to the ratio of the values in the columns of the $\mathrm{P}_{\text {skills }}$ matrix it is more convenient to rewrite the matrix into the form:

$$
P_{\text {skills }}=\begin{array}{l|ccc} 
& \text { Simple } & \text { Moderate } & \text { Complex } \\
\hline \text { Conscious } & \gamma_{1} p_{0} & \gamma_{2} p_{0} & \gamma_{3} p_{0} \\
\text { Typical } & * & * & * \\
\text { Naive } & 6 \gamma_{1} p_{0} & 6 \gamma_{2} p_{0} & 6 \gamma_{3} p_{0}
\end{array}
$$

Using these notations it is easy to express that the simulation studies covered the following sets of $\gamma$ parameters:

- $\gamma_{1}=6, \gamma_{2}=3, \gamma_{3}=1$,
- $\gamma_{1}=1, \gamma_{2}=3, \gamma_{3}=6$,
- $\gamma_{1}=1, \gamma_{2}=1, \gamma_{3}=1$.

Having investigated all three cases one can establish that the $\Delta(x)$ function is entirely insensitive to the different sets of $\gamma$ values. The details are not shown here.

### 7.4 Varying the ratio of the rows of $\mathrm{P}_{\text {skills }}$

To check the sensitivity of the relative improvement to the ratio of the values in the rows of the $P_{\text {skills }}$ matrix it is more convenient to rewrite the matrix into the form:

$$
\mathrm{P}_{\text {skills }}=\begin{array}{l|ccc} 
& \text { Simple } & \text { Moderate } & \text { Complex } \\
\hline \text { Conscious } & p_{0} & 3 p_{0} & 6 p_{0} \\
\text { Typical } & * & * & * \\
\text { Naive } & \delta p_{0} & 3 \delta p_{0} & 6 \delta p_{0}
\end{array}
$$

Using these notations it is easy to express that the simulation studies covered the following set of $\delta$ values: $\delta=1,2,4,6,12$. In Fig. 6 the $p(x)$ curves are shown for the "Naive-Conscious" "class change" situation for this set of $\delta$ values.


Figure 6: The $p(x)$ curve.

This sensitivity study when the effect of the varying $\delta$ value is being investigated is rather special. Since one specific $\delta$ value represents the differences between the Conscious and the Naive users in their skills (more precisely the ratio of the corresponding probabilities), it is straightforward that the effect of "reeducation" is larger if this difference is larger. If this difference is zero (the $\delta$ value is 1 ) the "reeducation" is obviously useless resulting in a $\Delta(x)=0$ value.
Hence this sensitivity study is mainly for learning the form of the curves describing the relationship between the number of "reeducated" users and the relative cyber security improvement and also learning the relationship between the difference of skills and the relative cyber security improvement rather than simply establishing the fact of the existence of this sensitivity for the varying $\delta$ values.

In Fig. 7(a) the almost perfect linear association can be established between the number of "reeducated" users ( $x$ ) and the relative cyber security improvement ( $\Delta(x)$ values) for all $\delta$ values.


Figure 7: Varying $\delta$.
In contrast the relationship between the difference in skills ( $\delta$ ) and the relative cyber security improvement $(\Delta(x))$ is far from linear. All curves for the different selected $x$ values show steep increase in the relative improvement as the value of $\delta$ starts to depart from its initial value of 1 but later the slopes of the curves are becoming smaller and smaller seemingly tending to a limit value of $\Delta(x)$.

From practical point of view the steep starting phase of the curves can be of prime interest. It means that even in case of small differences between the user groups' skills it may be worth to take measures for enhancing the users awareness since significant increase may happen in the level of cyber security.

## 8 Conclusion

In this paper the conditions and the findings of a simulation study was presented for assessing the effect size of users' consciousness to the computer network vulnerability in risky cyber attack situations at a certain business.

First a simple model was set up to classify the groups of users according to their skills and awareness then probabilities were assigned to each class describing the likelihood of committing dangerous reactions in case of a cyber attack.

To quantify the level of danger a metric developed in a former work was used. This $p_{s}$ metric shows the approximate probability of an infection at a given business with well specified parameters according to its location, the type of the attack, the protections used at the business etc.

To be able to see the tendencies in vulnerability exclusively attributable to the users consciousness the set of the numerous parameters were grouped to kept-constant and study parameters.

First the "typical study values" of the study parameters were used in the simulations then the sensitivity of the findings was investigated to the deviations from these typical values.

On one hand the findings mirrored back the straightforward and expected tendencies namely either the number of "reeducated" user is increasing or the surmounted difference of the users' groups in their skills is increasing the "relative improvement of the cyber security" is increasing.

On the other hand the tendencies in the change of this relative improvement have been established, different graphs and curves have been constructed to give an overall view for the influence of the different parameters.

In addition to these general conclusions assessments were made for the magnitude and for the range of the relative cyber security improvement. Slight sensitivity was experienced to the departures from the typical study values. It was shown that even in the case of small differences in skills making the users more conscious in their reactions can significantly enhance the level of cyber security at a business.

## References

[1] Bognár L, Joós A, Nagy B, An Improvement for a Mathematical Model for Distributed Vulnerability Assessment, Acta Universitatis Sapientiae, Mathematica, 10 (2) (2018), 203-217.
[2] Hadarics K, Györffy K, Nagy B, Bognár L, Arrott A and Leitold F, Mathematical Model of Distributed Vulnerability Assessment, In: Jaroslav Dočkal, Milan Jirsa, Josef Kaderka, Proceedings of Conference SPI 2017: Security and Protection of Information Brno, 2017.07.01-2017.07.02. Brno: University of Defence, 2017. pp. 45-57 (ISBN:978-80-7231-414-0).
[3] Leitold F, Arrott A and Hadarics K, Quantifying cyber-threat vulnerability by combining threat intelligence, IT infrastructure weakness, and user susceptibility 24th Annual EICAR Conference, Nuremberg, Germany, 2016.
[4] Leitold F, Hadarics K, Measuring security risk in the cloud-enabled enterprise In: Dr Fernando C Colon Osorio, 7th International Conference on Malicious and Unwanted Software (MALWARE), Fajardo, Puerto Rico, 2012.10.16-2012.10.18. Piscataway (NJ): IEEE, 2012. pp. 62-66 (ISBN:978-1-4673-4880-5).

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# Trapezoid type inequalities for generalized Riemann-Liouville fractional integrals of functions with bounded variation 

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#### Abstract

In this paper we establish some trapezoid type inequalities for the Riemann-Liouville fractional integrals of functions of bounded variation and of Hölder continuous functions. Applications for the gmean of two numbers are provided as well. Some particular cases for Hadamard fractional integrals are also provided.


## 1 Introduction

Let ( $\mathrm{a}, \mathrm{b}$ ) with $-\infty \leq \mathrm{a}<\mathrm{b} \leq \infty$ be a finite or infinite interval of the real line $\mathbb{R}$ and $\alpha$ a complex number with $\operatorname{Re}(\alpha)>0$. Also let $g$ be a strictly increasing function on $(a, b)$, having a continuous derivative $g^{\prime}$ on $(a, b)$. Following $[18$,

[^1]p. 100], we introduce the generalized left- and right-sided Riemann-Liouville fractional integrals of a function $f$ with respect to another function $g$ on $[a, b]$ by
$$
I_{a+, g}^{\alpha} f(x):=\frac{1}{\Gamma(\alpha)} \int_{a}^{x} \frac{g^{\prime}(t) f(t) d t}{[g(x)-g(t)]^{1-\alpha}}, a<x \leq b
$$
and
$$
I_{b-, g}^{\alpha} f(x):=\frac{1}{\Gamma(\alpha)} \int_{x}^{b} \frac{g^{\prime}(t) f(t) d t}{[g(t)-g(x)]^{1-\alpha}}, \quad a \leq x<b
$$

For $\mathrm{g}(\mathrm{t})=\mathrm{t}$ we have the classical Riemann-Liouville fractional integrals

$$
J_{a+}^{\alpha} f(x):=\frac{1}{\Gamma(\alpha)} \int_{a}^{x} \frac{f(t) d t}{(x-t)^{1-\alpha}}, a<x \leq b
$$

and

$$
J_{b-}^{\alpha} f(x):=\frac{1}{\Gamma(\alpha)} \int_{x}^{b} \frac{f(t) d t}{(t-x)^{1-\alpha}}, \quad a \leq x<b
$$

while for the logarithmic function $\mathrm{g}(\mathrm{t})=\ln \mathrm{t}$ we have the Hadamard fractional integrals [18, p. 111]

$$
H_{a+}^{\alpha} f(x):=\frac{1}{\Gamma(\alpha)} \int_{a}^{x}\left[\ln \left(\frac{x}{t}\right)\right]^{\alpha-1} \frac{f(t) d t}{t}, 0 \leq a<x \leq b
$$

and

$$
H_{b-}^{\alpha} f(x):=\frac{1}{\Gamma(\alpha)} \int_{x}^{b}\left[\ln \left(\frac{t}{x}\right)\right]^{\alpha-1} \frac{f(t) d t}{t}, 0 \leq a<x<b
$$

One can consider the function $g(t)=-t^{-1}$ and define the "Harmonic fractional integrals" by

$$
R_{a+}^{\alpha} f(x):=\frac{x^{1-\alpha}}{\Gamma(\alpha)} \int_{a}^{x} \frac{f(t) d t}{(x-t)^{1-\alpha} t^{\alpha+1}}, 0 \leq a<x \leq b
$$

and

$$
\mathrm{R}_{\mathrm{b}-}^{\alpha} f(x):=\frac{x^{1-\alpha}}{\Gamma(\alpha)} \int_{x}^{b} \frac{f(t) d t}{(t-x)^{1-\alpha} t^{\alpha+1}}, 0 \leq a<x<b
$$

Also, for $g(t)=\exp (\beta t), \beta>0$, we can consider the " $\beta$-Exponential fractional integrals"

$$
E_{a+, \beta}^{\alpha} f(x):=\frac{\beta}{\Gamma(\alpha)} \int_{a}^{x} \frac{\exp (\beta t) f(t) d t}{[\exp (\beta x)-\exp (\beta t)]^{1-\alpha}}, a<x \leq b
$$

and

$$
E_{b-, \beta}^{\alpha} f(x):=\frac{\beta}{\Gamma(\alpha)} \int_{x}^{b} \frac{\exp (\beta t) f(t) d t}{[\exp (\beta t)-\exp (\beta x)]^{1-\alpha}}, a \leq x<b
$$

In the recent paper [14] we obtained the following Ostrowski type inequalities for functions of bounded variation:

Theorem 1 Let $f:[a, b] \rightarrow \mathbb{C}$ be a function of bounded variation on $[\mathrm{a}, \mathrm{b}]$ and g be a strictly increasing function on $(\mathrm{a}, \mathrm{b})$, having a continuous derivative $\mathrm{g}^{\prime}$ on $(\mathrm{a}, \mathrm{b})$. For any $\mathrm{x} \in(\mathrm{a}, \mathrm{b})$ we have the inequalities

$$
\left.\begin{array}{l}
\left\lvert\, \begin{array}{rl}
\mid I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)- & \left.\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right) f(x) \right\rvert\, \\
\leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t) \bigvee_{t}^{x}(f) d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t) \bigvee_{x}^{t}(f) d t}{[g(t)-g(x)]^{1-\alpha}}\right]
\end{array}\right. \\
\leq \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha} \bigvee_{a}^{x}(f)+[g(b)-g(x)]^{\alpha} \bigvee_{x}^{b}(f)\right]
\end{array}\right] \begin{aligned}
& {\left[\frac{1}{2}(g(b)-g(a))+\left|g(x)-\frac{g(a)+g(b)}{2}\right|\right]^{\alpha} \bigvee_{a}^{b}(f) ;} \\
& \leq \frac{1}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
\left((g(x)-g(a))^{\alpha p}+(g(b)-g(x))^{\alpha p}\right)^{1 / p}\left(\left(V_{a}^{x}(f)\right)^{q}+\left(V_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
\text { with p,q>1, } \frac{1}{p}+\frac{1}{q}=1 ; \\
\left((g(x)-g(a))^{\alpha}+(g(b)-g(x))^{\alpha}\right)\left[\frac{1}{2} V_{a}^{b}(f)+\frac{1}{2}\left|V_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{array}\right.
\end{aligned}
$$

and

$$
\left.\left.\begin{array}{rl}
\mid I_{x-, g}^{\alpha} f(a)+ & \left.I_{x+, g}^{\alpha} f(b)-\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right) f(x) \right\rvert\, \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t) \bigvee_{t}^{x}(f) d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t) \bigvee_{x}^{t}(f) d t}{[g(b)-g(t)]^{1-\alpha}}\right] \\
\leq & \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha} \bigvee_{a}^{x}(f)+[g(b)-g(x)]^{\alpha} \bigvee_{x}^{b}(f)\right]
\end{array}\right] \begin{array}{l}
{\left[\frac{1}{2}(g(b)-g(a))+\left|g(x)-\frac{g(a)+g(b)}{2}\right|\right]^{\alpha} V_{a}^{b}(f) ;}
\end{array}\right\} \begin{aligned}
& \left((g(x)-g(a))^{\alpha p}+(g(b)-g(x))^{\alpha p}\right)^{1 / p}\left(\left(V_{a}^{x}(f)\right)^{q}+\left(V_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
& \text { with p,q>1, } \frac{1}{p}+\frac{1}{q}=1 ; \\
& \left((g(x)-g(a))^{\alpha}+(g(b)-g(x))^{\alpha}\right)\left[\frac{1}{2} \bigvee_{a}^{b}(f)+\frac{1}{2}\left|V_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{aligned}
$$

If g is a function which maps an interval I of the real line to the real numbers, and is both continuous and injective then we can define the g-mean of two numbers $\mathrm{a}, \mathrm{b} \in \mathrm{I}$ as

$$
M_{g}(a, b):=g^{-1}\left(\frac{g(a)+g(b)}{2}\right)
$$

If $I=\mathbb{R}$ and $g(t)=t$ is the identity function, then $M_{g}(a, b)=A(a, b):=$ $\frac{a+b}{2}$, the arithmetic mean. If $I=(0, \infty)$ and $g(t)=\ln t$, then $M_{g}(a, b)=$ $\mathrm{G}(\mathrm{a}, \mathrm{b}):=\sqrt{\mathrm{ab}}$, the geometric mean. If $\mathrm{I}=(0, \infty)$ and $g(t)=\frac{1}{\mathrm{t}}$, then $M_{g}(a, b)=H(a, b):=\frac{2 a b}{a+b}$, the harmonic mean. If $I=(0, \infty)$ and $g(t)=t^{p}$, $p \neq 0$, then $M_{g}(a, b)=M_{p}(a, b):=\left(\frac{a^{p}+b^{p}}{2}\right)^{1 / p}$, the power mean with exponent $p$. Finally, if $I=\mathbb{R}$ and $g(t)=\exp t$, then

$$
M_{g}(a, b)=\operatorname{LME}(a, b):=\ln \left(\frac{\exp a+\exp b}{2}\right)
$$

the LogMeanExp function.
The following particular case for g-mean is of interest [14].
Corollary 1 With the assumptions of Theorem 1 we have

$$
\begin{gathered}
\left|I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right)-\frac{[g(b)-g(a)]^{\alpha}}{2^{\alpha-1} \Gamma(\alpha+1)} f\left(M_{g}(a, b)\right)\right| \\
\leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t) \bigvee_{t}^{M_{g}(a, b)}(f) d t}{\left[g\left(M_{g}(a, b)\right)-g(t)\right]^{1-\alpha}}+\int_{M_{g}(a, b)\left[g(t)-g\left(M_{g}(a, b)\right)\right]^{1-\alpha}}^{b} \frac{g^{\prime}(t) \bigvee_{M_{g}(a, b)}^{t}(f) d t}{[g}\right] \\
\leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}(g(b)-g(a))^{\alpha} \bigvee_{a}^{b}(f)
\end{gathered}
$$

and

$$
\begin{aligned}
& \left|I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b)-\frac{[g(b)-g(a)]^{\alpha}}{2^{\alpha-1} \Gamma(\alpha+1)} f\left(M_{g}(a, b)\right)\right| \\
& \quad \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t) \bigvee_{t}^{M_{g}(a, b)}(f) d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t) \bigvee_{x}^{t}(f) d t}{[g(b)-g(t)]^{1-\alpha}}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}(g(b)-g(a))^{\alpha} \bigvee_{a}^{b}(f)
\end{aligned}
$$

Remark 1 If we take in Theorem $1 \mathrm{x}=\frac{\mathrm{a}+\mathrm{b}}{2}$, then we obtain similar mid-point inequalities, however the details are not presented here. Some applications for the Hadamard fractional integrals are also provided in [14].

For several Ostrowski type inequalities for Riemann-Liouville fractional integrals see [1]-[5], [16]-[27] and the references therein.

Motivated by the above results, in this paper we establish some trapezoid type inequalities for the generalized Riemann-Liouville fractional integrals of functions of bounded variation and of Hölder continuous functions. Applications for the g-mean of two numbers are provided as well. Some particular cases for Hadamard fractional integrals are also provided.

## 2 Some identities

We have:
Lemma 1 Let $\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{C}$ be Lebesgue integrable on $[\mathrm{a}, \mathrm{b}], \mathrm{g}$ be a strictly increasing function on $(\mathrm{a}, \mathrm{b})$, having a continuous derivative $\mathrm{g}^{\prime}$ on $(\mathrm{a}, \mathrm{b})$ and $\lambda, \mu$ some complex parameters:
(i) For any $\mathrm{x} \in(\mathrm{a}, \mathrm{b})$ we have the representation

$$
\begin{array}{r}
I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)=\frac{1}{\Gamma(\alpha+1)}\left(\lambda[g(x)-g(a)]^{\alpha}+\mu[g(b)-g(x)]^{\alpha}\right) \\
+\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(t)-g(x)]^{1-\alpha}}\right] \tag{1}
\end{array}
$$

and

$$
\begin{align*}
I_{x-, g}^{\alpha} f(a)+I_{x+, g}^{\alpha} f(b) & =\frac{1}{\Gamma(\alpha+1)}\left(\lambda[g(x)-g(a)]^{\alpha}+\mu[g(b)-g(x)]^{\alpha}\right) \\
& +\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{\alpha} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(b)-g(t)]^{1-\alpha}}\right] \tag{2}
\end{align*}
$$

(ii) We have

$$
\begin{align*}
& \frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}=\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \frac{\lambda+\mu}{2} \\
&+\frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(t)-g(a)]^{1-\alpha}}\right] \tag{3}
\end{align*}
$$

Proof. (i) We observe that

$$
\begin{align*}
& \frac{1}{\Gamma(\alpha)} \int_{a}^{x} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(x)-g(t)]^{1-\alpha}}  \tag{4}\\
& =I_{a+, g}^{\alpha} f(x)-\lambda \frac{1}{\Gamma(\alpha)} \int_{a}^{x} \frac{g^{\prime}(t) d t}{[g(x)-g(t)]^{1-\alpha}} \\
& =I_{a+, g}^{\alpha} f(x)-\frac{[g(x)-g(a)]^{\alpha}}{\alpha \Gamma(\alpha)} \lambda=I_{a+, g}^{\alpha} f(x)-\frac{[g(x)-g(a)]^{\alpha}}{\Gamma(\alpha+1)} \lambda
\end{align*}
$$

for $\mathrm{a}<\mathrm{x} \leq \mathrm{b}$ and, similarly,

$$
\begin{equation*}
\frac{1}{\Gamma(\alpha)} \int_{x}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(t)-g(x)]^{1-\alpha}}=I_{b-, g}^{\alpha} f(x)-\frac{[g(b)-g(x)]^{\alpha}}{\Gamma(\alpha+1)} \mu \tag{5}
\end{equation*}
$$

for $a \leq x<b$.
If $x \in(a, b)$, then by adding the equalities (4) and (5) we get the representation (1).

By the definition of fractional integrals we have

$$
I_{x+, g}^{\alpha} f(b):=\frac{1}{\Gamma(\alpha)} \int_{x}^{b} \frac{g^{\prime}(t) f(t) d t}{[g(b)-g(t)]^{1-\alpha}}, \quad a \leq x<b
$$

and

$$
I_{x-, g}^{\alpha} f(a):=\frac{1}{\Gamma(\alpha)} \int_{a}^{x} \frac{g^{\prime}(t) f(t) d t}{[g(t)-g(a)]^{1-\alpha}}, \quad a<x \leq b
$$

Then

$$
\begin{equation*}
\frac{1}{\Gamma(\alpha)} \int_{x}^{b} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(b)-g(t)]^{1-\alpha}}=I_{x+, g}^{\alpha} f(b)-\frac{[g(b)-g(x)]^{\alpha}}{\Gamma(\alpha+1)} \lambda \tag{6}
\end{equation*}
$$

for $a \leq x<b$ and

$$
\begin{equation*}
\frac{1}{\Gamma(\alpha)} \int_{a}^{\alpha} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(t)-g(a)]^{1-\alpha}}=I_{x-, g}^{\alpha} f(a)-\frac{[g(x)-g(a)]^{\alpha}}{\Gamma(\alpha+1)} \mu \tag{7}
\end{equation*}
$$

for $a<x \leq b$.
If $x \in(a, b)$, then by adding the equalities (6) and (7) we get the representation (1).

If we take $x=b$ in (4) we get

$$
\begin{equation*}
\frac{1}{\Gamma(\alpha)} \int_{a}^{b} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(b)-g(t)]^{1-\alpha}}=I_{a+, g}^{\alpha} f(b)-\frac{[g(b)-g(a)]^{\alpha}}{\Gamma(\alpha+1)} \lambda \tag{8}
\end{equation*}
$$

while from $x=a$ in (5) we get

$$
\begin{equation*}
\frac{1}{\Gamma(\alpha)} \int_{a}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(t)-g(a)]^{1-\alpha}}=I_{b-, g}^{\alpha} f(a)-\frac{[g(b)-g(a)]^{\alpha}}{\Gamma(\alpha+1)} \mu . \tag{9}
\end{equation*}
$$

If we add (8) with (9) and divide by 2 we get (3).
Remark 2 If we take in (1) and (2) $x=M_{g}(a, b)=g^{-1}\left(\frac{g(a)+g(b)}{2}\right)$, then we get

$$
\begin{gathered}
I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right) \\
=\frac{1}{2^{\alpha-1} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\left(\frac{\lambda+\mu}{2}\right) \\
+\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{\left[g\left(M_{g}(a, b)\right)-g(t)\right]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{\left[g(t)-g\left(M_{g}(a, b)\right)\right]^{1-\alpha}}\right]
\end{gathered}
$$

and

$$
\begin{aligned}
& I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b)=\frac{1}{2^{\alpha-1} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\left(\frac{\lambda+\mu}{2}\right) \\
& \quad+\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t)[f(t)-\lambda] d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t)[f(t)-\mu] d t}{[g(b)-g(t)]^{1-\alpha}}\right]
\end{aligned}
$$

The above lemma provides various identities of interest by taking particular values for the parameters $\lambda$ and $\mu$, out of which we give only a few:

Corollary 2 With the assumptions of Lemma 1 we have:
(i) For any $x \in(a, b)$,

$$
\begin{align*}
& I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)=\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right) f(x) \\
& \quad+\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)[f(t)-f(x)] d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)[f(t)-f(x)] d t}{[g(t)-g(x)]^{1-\alpha}}\right] \tag{10}
\end{align*}
$$

and

$$
I_{x-, g}^{\alpha} f(a)+I_{x+, g}^{\alpha} f(b)=\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right) f(x)
$$

$$
\begin{equation*}
+\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)[f(t)-f(x)] d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)[f(t)-f(x)] d t}{[g(b)-g(t)]^{1-\alpha}}\right] \tag{11}
\end{equation*}
$$

(ii) For any $x \in[a, b]$,

$$
\begin{align*}
& \frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}=\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} f(x) \\
& \quad+\frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t)[f(t)-f(x)] d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t)[f(t)-f(x)] d t}{[g(t)-g(a)]^{1-\alpha}}\right] \tag{12}
\end{align*}
$$

The proof is obvious by taking $\lambda=\mu=\mathrm{f}(\mathrm{x})$ in Lemma 1. These identities were obtained in [14]. If we take in (10)-(12) $x=M_{g}(a, b)=g^{-1}\left(\frac{g(a)+g(b)}{2}\right)$, then we get the corresponding identities were obtained in [14].

Corollary 3 With the assumptions of Lemma 1 we have:

$$
\begin{align*}
I_{a+, g}^{\alpha} f(x) & +I_{b-, g}^{\alpha} f(x) \\
& =\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)\right) \\
+ & \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)[f(t)-f(a)] d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)[f(t)-f(b)] d t}{[g(t)-g(x)]^{1-\alpha}}\right] \tag{13}
\end{align*}
$$

and

$$
\begin{align*}
I_{x-, g}^{\alpha} f(a) & +I_{x+, g}^{\alpha} f(b) \\
& =\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)\right) \\
+ & \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)[f(t)-f(a)] d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)[f(t)-f(b)] d t}{[g(b)-g(t)]^{1-\alpha}}\right] \tag{14}
\end{align*}
$$

for any $x \in(a, b)$
(ii) We also have

$$
\begin{align*}
& \frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}=\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \frac{f(b)+f(a)}{2} \\
& \quad+\frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t)[f(t)-f(b)] d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t)[f(t)-f(a)] d t}{[g(t)-g(a)]^{1-\alpha}}\right] . \tag{15}
\end{align*}
$$

The proof of (13) and (14) are obvious by taking $\lambda=f(a), \mu=f(b)$ in Lemma 1. The proof of (15) follows by Lemma 1 on taking $\lambda=f(b)$ and $\mu=\mathrm{f}(\mathrm{a})$.

Remark 3 If we take in (13) and (14) $x=M_{g}(a, b)=g^{-1}\left(\frac{g(a)+g(b)}{2}\right)$, then we get

$$
\begin{gathered}
I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right) \\
=\frac{1}{2^{\alpha-1} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\left(\frac{f(a)+f(b)}{2}\right) \\
+\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t)[f(t)-f(a)] d t}{\left[g\left(M_{g}(a, b)\right)-g(t)\right]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t)[f(t)-f(b)] d t}{\left[g(t)-g\left(M_{g}(a, b)\right)\right]^{1-\alpha}}\right]
\end{gathered}
$$

and

$$
\begin{aligned}
& I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b) \\
& \quad=\frac{1}{2^{\alpha-1} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\left(\frac{f(a)+f(b)}{2}\right) \\
& +\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t)[f(t)-f(a)] d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t)[f(t)-f(b)] d t}{[g(b)-g(t)]^{1-\alpha}}\right]
\end{aligned}
$$

## 3 Inequalities for bounded functions

Now, for $\phi, \Phi \in \mathbb{C}$ and $[a, b]$ an interval of real numbers, define the sets of complex-valued functions, see for instance [15]
$\overline{\mathrm{U}}_{[\mathrm{a}, \mathrm{b}]}(\phi, \Phi)$
$:=\{\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{C} \mid \operatorname{Re}[(\Phi-\mathrm{f}(\mathrm{t}))(\overline{\mathrm{f}(\mathrm{t})}-\bar{\phi})] \geq 0$ for almost every $\mathrm{t} \in[\mathrm{a}, \mathrm{b}]\}$
and
$\bar{\Delta}_{[a, b]}(\phi, \Phi):=\left\{f: \left.[a, b] \rightarrow \mathbb{C}| | f(t)-\frac{\phi+\Phi}{2}\left|\leq \frac{1}{2}\right| \Phi-\phi \right\rvert\,\right.$ for a.e. $\left.t \in[a, b]\right\}$.
The following representation result may be stated.
Proposition 1 For any $\phi, \Phi \in \mathbb{C}, \phi \neq \Phi$, we have that $\overline{\mathrm{u}}_{[\mathrm{a}, \mathrm{b}]}(\phi, \Phi)$ and $\bar{\Delta}_{[\mathrm{a}, \mathrm{b}]}(\phi, \Phi)$ are nonempty, convex and closed sets and

$$
\begin{equation*}
\overline{\mathrm{u}}_{[\mathrm{a}, \mathrm{~b}]}(\phi, \Phi)=\bar{\Delta}_{[\mathrm{a}, \mathrm{~b}]}(\phi, \Phi) \tag{16}
\end{equation*}
$$

Proof. We observe that for any $z \in \mathbb{C}$ we have the equivalence

$$
\left|z-\frac{\phi+\Phi}{2}\right| \leq \frac{1}{2}|\Phi-\phi|
$$

if and only if

$$
\operatorname{Re}[(\Phi-z)(\bar{z}-\phi)] \geq 0
$$

This follows by the equality

$$
\frac{1}{4}|\Phi-\phi|^{2}-\left|z-\frac{\phi+\Phi}{2}\right|^{2}=\operatorname{Re}[(\Phi-z)(\bar{z}-\phi)]
$$

that holds for any $z \in \mathbb{C}$.
The equality (16) is thus a simple consequence of this fact.
On making use of the complex numbers field properties we can also state that:

Corollary 4 For any $\phi, \Phi \in \mathbb{C}, \phi \neq \Phi$, we have that

$$
\begin{aligned}
\overline{\mathrm{U}}_{[\mathrm{a}, \mathrm{~b}]}(\phi, \Phi)= & \{\mathrm{f}:[\mathrm{a}, \mathrm{~b}] \rightarrow \mathbb{C} \mid(\operatorname{Re} \Phi-\operatorname{Ref}(\mathrm{t}))(\operatorname{Ref}(\mathrm{t})-\operatorname{Re} \phi) \\
& +(\operatorname{Im} \Phi-\operatorname{Imf}(\mathrm{t}))(\operatorname{Imf}(\mathrm{t})-\operatorname{Im} \phi) \geq 0 \text { for a.e. } \mathrm{t} \in[\mathrm{a}, \mathrm{~b}]\} .
\end{aligned}
$$

Now, if we assume that $\operatorname{Re}(\Phi) \geq \operatorname{Re}(\phi)$ and $\operatorname{Im}(\Phi) \geq \operatorname{Im}(\phi)$, then we can define the following set of functions as well:

$$
\begin{aligned}
\bar{S}_{[a, b]}(\phi, \Phi) & :=\{\mathrm{f}:[\mathrm{a}, \mathrm{~b}] \rightarrow \mathbb{C} \mid \operatorname{Re}(\Phi) \geq \operatorname{Ref}(\mathrm{t}) \geq \operatorname{Re}(\phi) \\
& \text { and } \operatorname{Im}(\Phi) \geq \operatorname{Imf}(\mathrm{t}) \geq \operatorname{Im}(\phi) \text { for a.e. } \mathrm{t} \in[\mathrm{a}, \mathrm{~b}]\}
\end{aligned}
$$

One can easily observe that $\bar{S}_{[a, b]}(\phi, \Phi)$ is closed, convex and

$$
\emptyset \neq \overline{\mathrm{S}}_{[\mathrm{a}, \mathrm{~b}]}(\phi, \Phi) \subseteq \overline{\mathrm{U}}_{[\mathrm{a}, \mathrm{~b}]}(\phi, \Phi)
$$

We have:
Theorem 2 Let $\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{C}$ be a complex valued Lebesgue integrable function on the real interval $[\mathrm{a}, \mathrm{b}], \mathrm{g}$ be a strictly increasing function on $(\mathrm{a}, \mathrm{b})$, having a continuous derivative $\mathrm{g}^{\prime}$ on $(\mathrm{a}, \mathrm{b})$ and $\phi, \Phi \in \mathbb{C}, \phi \neq \Phi$ such that $f \in \bar{\Delta}_{[a, b]}(\phi, \Phi)$.
(i) For any $x \in(a, b)$,

$$
\begin{equation*}
\left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{\phi+\Phi}{2 \Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right)\right| \tag{17}
\end{equation*}
$$

$$
\leq \frac{1}{2}|\Phi-\phi| \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right]
$$

and

$$
\begin{align*}
& \left|I_{x-, g}^{\alpha} f(a)+I_{x+, g}^{\alpha} f(b)-\frac{\phi+\Phi}{2 \Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right)\right|  \tag{18}\\
& \leq \frac{1}{2}|\Phi-\phi| \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right] .
\end{align*}
$$

(ii) We have

$$
\begin{align*}
& \left|\frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}-\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \frac{\phi+\Phi}{2}\right|  \tag{19}\\
& \leq \frac{1}{2}|\Phi-\phi| \frac{1}{\Gamma(\alpha+1)}|\Phi-\phi|[g(b)-g(a)]^{\alpha} .
\end{align*}
$$

Proof. Using the identity (1) for $\lambda=\mu=\frac{\phi+\Phi}{2}$, we have

$$
\begin{align*}
& \mathrm{I}_{a+, \mathrm{g}}^{\alpha} \mathrm{f}(\mathrm{x})+\mathrm{I}_{\mathrm{b}-, \mathrm{g}}^{\alpha} \mathrm{f}(\mathrm{x}) \\
& \quad-\frac{1}{\Gamma(\alpha+1)}\left([\mathrm{g}(\mathrm{x})-\mathrm{g}(\mathrm{a})]^{\alpha}+[\mathrm{g}(\mathrm{~b})-\mathrm{g}(\mathrm{x})]^{\alpha}\right) \frac{\phi+\Phi}{2} \\
& =\frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(\mathrm{t})\left[f(\mathrm{f})-\frac{\phi+\Phi}{2}\right] d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)\left[f(t)-\frac{\phi+\Phi}{2}\right] d t}{[g(t)-g(x)]^{1-\alpha}}\right] \tag{20}
\end{align*}
$$

for any $x \in(a, b)$.
Taking the modulus in (20), then we get

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{1}{\Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right) \frac{\phi+\Phi}{2}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)\left|f(t)-\frac{\phi+\Phi}{2}\right| d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)\left|f(t)-\frac{\phi+\Phi}{2}\right| d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
& \leq \frac{1}{2}|\Phi-\phi| \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t) d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t) d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
& =\frac{1}{2}|\Phi-\phi| \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right]
\end{aligned}
$$

for any $x \in(a, b)$, which proves (17).

The inequality (18) follows in a similar manner from the identity (2).
The inequality (19) follows by (3) for $\lambda=\mu=\frac{\phi+\Phi}{2}$.
Corollary 5 With the assumptions of Theorem 2 we have

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right)-\frac{\phi+\Phi}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{1}{2^{\alpha}}|\Phi-\phi| \frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b)-\frac{\phi+\Phi}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{1}{2^{\alpha}}|\Phi-\phi| \frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} .
\end{aligned}
$$

Remark 4 If the function $\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{R}$ is measurable and there exists the constants $\mathrm{m}, \mathrm{M}$ such that $\mathrm{m} \leq \mathrm{f}(\mathrm{t}) \leq M$ for a.e. $\mathrm{t} \in[\mathrm{a}, \mathrm{b}]$, then for any $x \in(a, b)$ we have by (17) and (18) that

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{m+M}{2 \Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right)\right| \\
& \leq \frac{1}{2}(M-m) \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right]
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|I_{x-, g}^{\alpha} f(a)+I_{x+, g}^{\alpha} f(b)-\frac{m+M}{2 \Gamma(\alpha+1)}\left([g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right)\right| \\
& \leq \frac{1}{2}(M-m) \frac{1}{\Gamma(\alpha+1)}\left[[g(x)-g(a)]^{\alpha}+[g(b)-g(x)]^{\alpha}\right]
\end{aligned}
$$

In particular,

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right)-\frac{m+M}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{1}{2^{\alpha}}(M-m) \frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b)-\frac{m+M}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{1}{2^{\alpha}}(M-m) \frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}
\end{aligned}
$$

## 4 Trapezoid inequalities for functions of bounded variation

We have:
Theorem 3 Let $\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{C}$ be a complex valued function of bounded variation on the real interval $[\mathrm{a}, \mathrm{b}]$, and g be a strictly increasing function on $(\mathrm{a}, \mathrm{b})$, having a continuous derivative $\mathrm{g}^{\prime}$ on $(\mathrm{a}, \mathrm{b})$. Then we have the inequalities

$$
\begin{align*}
& \left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{[g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
& \leq \frac{1}{\Gamma(\alpha+1)}\left[(g(x)-g(a))^{\alpha} \bigvee_{a}^{x}(f)+(g(b)-g(x))^{\alpha} \bigvee_{x}^{b}(f)\right] \\
& \int\left[\frac{1}{2}(g(b)-g(a))+\left|g(x)-\frac{g(a)+g(b)}{2}\right|\right]^{\alpha} V_{a}^{b}(f) ; \\
& \leq \frac{1}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
\left((g(x)-g(a))^{\alpha p}+(g(b)-g(x))^{\alpha p}\right)^{1 / p}\left(\left(\bigvee_{a}^{x}(f)\right)^{q}+\left(\bigvee_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
\text { with p,q>1, } \frac{1}{p}+\frac{1}{q}=1 ; \\
\left((g(x)-g(a))^{\alpha}+(g(b)-g(x))^{\alpha}\right)\left[\frac{1}{2} \bigvee_{a}^{b}(f)+\frac{1}{2}\left|\bigvee_{a}^{x}(f)-\bigvee_{x}^{b}(f)\right|\right]
\end{array}\right. \tag{21}
\end{align*}
$$

and

$$
\left.\left.\begin{array}{l}
\left|I_{x-, g}^{\alpha} f(a)+I_{x+, g}^{\alpha} f(b)-\frac{[g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
\leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{[g(b)-g(t)]^{1-\alpha}}\right] \\
\leq \frac{1}{\Gamma(\alpha+1)}\left[(g(x)-g(a))^{\alpha} \bigvee_{a}^{x}(f)+(g(b)-g(x))^{\alpha} \bigvee_{x}^{b}(f)\right]
\end{array}\right] \begin{array}{l}
{\left[\frac{1}{2}(g(b)-g(a))+\left|g(x)-\frac{g(a)+g(b)}{2}\right|\right]^{\alpha} V_{a}^{b}(f) ;} \\
\leq \frac{1}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
\left((g(x)-g(a))^{\alpha p}+(g(b)-g(x))^{\alpha p}\right)^{1 / p}\left(\left(V_{a}^{x}(f)\right)^{q}+\left(V_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
\left(\left(g i t h p, q>1, \frac{1}{p}+\frac{1}{q}=1 ;\right.\right.
\end{array}\right.  \tag{22}\\
\left.(x)-g(a))^{\alpha}+(g(b)-g(x))^{\alpha}\right)\left[\frac{1}{2} \bigvee_{a}^{b}(f)+\frac{1}{2}\left|V_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{array}\right] .
$$

for any $x \in(a, b)$
(ii) We also have

$$
\begin{align*}
& \left|\frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}-\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \frac{f(b)+f(a)}{2}\right| \\
& \leq \frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{[g(t)-g(a)]^{1-\alpha}}\right]  \tag{23}\\
& \leq \frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \bigvee_{a}^{b}(f) .
\end{align*}
$$

Proof. Using the identity (13) and the properties of the modulus, we have

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{[g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)|f(t)-f(a)| d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)|f(t)-f(b)| d t}{[g(t)-g(x)]^{1-\alpha}}\right]=: B(x)
\end{aligned}
$$

for any $x \in(a, b)$.
Since $f$ is of bounded variation on $[a, b]$, then we have

$$
|f(t)-f(a)| \leq \bigvee_{a}^{t}(f) \leq \bigvee_{a}^{x}(f) \text { for } a \leq t \leq x
$$

and

$$
|f(t)-f(b)| \leq \bigvee_{t}^{b}(f) \leq \bigvee_{x}^{b}(f) \text { for } x \leq t \leq b
$$

Therefore

$$
\begin{aligned}
B(x) & \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\bigvee_{a}^{x}(f) \int_{a}^{x} \frac{g^{\prime}(t) d t}{[g(x)-g(t)]^{1-\alpha}}+\bigvee_{x}^{b}(f) \int_{x}^{b} \frac{g^{\prime}(t) d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
& =\frac{1}{\Gamma(\alpha)}\left[\frac{(g(x)-g(a))^{\alpha}}{\alpha} \bigvee_{a}^{x}(f)+\frac{(g(b)-g(x))^{\alpha}}{\alpha} \bigvee_{x}^{b}(f)\right] \\
& =\frac{1}{\Gamma(\alpha+1)}\left[(g(x)-g(a))^{\alpha} \bigvee_{a}^{x}(f)+(g(b)-g(x))^{\alpha} \bigvee_{x}^{b}(f)\right]
\end{aligned}
$$

which proves the first two inequalities in (21).
The last part of (21) is obvious by making use of the elementary Hölder type inequalities for positive real numbers $c, d, m, n \geq 0$

$$
m c+n d \leq\left\{\begin{array}{l}
\max \{m, n\}(c+d) \\
\left(m^{p}+n^{p}\right)^{1 / p}\left(c^{q}+d^{q}\right)^{1 / q} \text { with } p, q>1, \frac{1}{p}+\frac{1}{q}=1
\end{array}\right.
$$

The inequality (22) follows in a similar way by utilising the equality (14).
From the equality (15) we have

$$
\begin{aligned}
& \left|\frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}-\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \frac{f(b)+f(a)}{2}\right| \\
& \leq \frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t)|f(t)-f(b)| d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t)|f(t)-f(a)| d t}{[g(t)-g(a)]^{1-\alpha}}\right] \\
& \leq \frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{[g(t)-g(a)]^{1-\alpha}}\right] \\
& \leq \frac{1}{2 \Gamma(\alpha)}\left[\bigvee_{a}^{b}(f) \int_{a}^{b} \frac{g^{\prime}(t) d t}{[g(b)-g(t)]^{1-\alpha}}+\bigvee_{a}^{b}(f) \int_{a}^{b} \frac{g^{\prime}(t) d t}{[g(t)-g(a)]^{1-\alpha}}\right] \\
& =\frac{1}{2 \Gamma(\alpha)}\left[\bigvee_{a}^{b}(f) \frac{[g(b)-g(a)]^{\alpha}}{\alpha}+\bigvee_{a}^{b}(f) \frac{[g(b)-g(a)]^{\alpha}}{\alpha}\right] \\
& =\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \bigvee_{a}^{b}(f),
\end{aligned}
$$

which proves (23).

Corollary 6 With the assumptions of Theorem 3 we have

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right)-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{\left[g\left(M_{g}(a, b)\right)-g(t)\right]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{\left[g(t)-g\left(M_{g}(a, b)\right)\right]^{1-\alpha}}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}(g(b)-g(a))^{\alpha} \bigvee_{a}^{b}(f)
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b)-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t) \bigvee_{t}^{b}(f) d t}{[g(b)-g(t)]^{1-\alpha}}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}(g(b)-g(a))^{\alpha} \bigvee_{a}^{b}(f) .
\end{aligned}
$$

## 5 Inequalities for Hölder's continuous functions

We say that the function $\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{C}$ is r - H -Hölder continuous on $[\mathrm{a}, \mathrm{b}]$ with $r \in(0,1]$ and $H>0$ if

$$
\begin{equation*}
|f(t)-f(s)| \leq H|t-s|^{r} \tag{24}
\end{equation*}
$$

for any $t, s \in[a, b]$. If $r=1$ and $H=L$ we call the function $L$-Lipschitzian on [a,b].

Theorem 4 Assume that $\mathrm{f}:[\mathrm{a}, \mathrm{b}] \rightarrow \mathbb{C}$ is r -H-Hölder continuous on $[\mathrm{a}, \mathrm{b}]$ with $\mathrm{r} \in(0,1]$ and $\mathrm{H}>0$, and g be a strictly increasing function on $(\mathrm{a}, \mathrm{b})$, having a continuous derivative $\mathrm{g}^{\prime}$ on $(\mathrm{a}, \mathrm{b})$. Then

$$
\left.\begin{array}{l}
\left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{[g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
\leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)(t-a)^{r} d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)(b-t)^{r} d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
\leq \frac{H}{\Gamma(\alpha+1)}\left[(g(x)-g(a))^{\alpha}(x-a)^{r}+(g(b)-g(x))^{\alpha}(b-x)^{r}\right]
\end{array}\right\} \begin{aligned}
& {\left[\begin{array}{l}
{\left[\frac{1}{2}(g(b)-g(a))+\left|g(x)-\frac{g(a)+g(b)}{2}\right|\right]^{\alpha}\left[(x-a)^{r}+(b-x)^{r}\right]}
\end{array}\right.} \\
& \leq \frac{H}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
\left((g(x)-g(a))^{\alpha p}+(g(b)-g(x))^{\alpha p}\right)^{1 / p}\left((x-a)^{r q}+(b-x)^{r q}\right)^{1 / q} \\
w i t h p, q>1, \frac{1}{p}+\frac{1}{q}=1 ; \\
\left((g(x)-g(a))^{\alpha}+(g(b)-g(x))^{\alpha}\right)\left[\frac{1}{2}(b-a)+\left|x-\frac{a+b}{2}\right|^{r}\right]^{r}
\end{array}\right. \tag{25}
\end{aligned}
$$

and

$$
\left.\begin{array}{l}
\left.I_{x-, g}^{\alpha} f(a)+I_{x+, g}^{\alpha} f(b)-\frac{[g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)}{\Gamma(\alpha+1)} \right\rvert\, \\
\leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)(t-a)^{r} d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)(b-t)^{r} d t}{[g(b)-g(t)]^{1-\alpha}}\right] \\
\leq \frac{H}{\Gamma(\alpha+1)}\left[(g(x)-g(a))^{\alpha}(x-a)^{r}+(g(b)-g(x))^{\alpha}(b-x)^{r}\right]
\end{array}\right\} \begin{aligned}
& {\left[\begin{array}{l}
{\left[\frac{1}{2}(g(b)-g(a))+\left|g(x)-\frac{g(a)+g(b)}{2}\right|\right]^{\alpha}\left[(x-a)^{r}+(b-x)^{r}\right]} \\
\leq \frac{H}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
\left((g(x)-g(a))^{\alpha p}+(g(b)-g(x))^{\alpha p}\right)^{1 / p}\left((x-a)^{r q}+(b-x)^{r q}\right)^{1 / q} \\
w i t h p, q>1, \frac{1}{p}+\frac{1}{q}=1 ; \\
\left((g(x)-g(a))^{\alpha}+(g(b)-g(x))^{\alpha}\right)\left[\frac{1}{2}(b-a)+\left|x-\frac{a+b}{2}\right|^{r}\right.
\end{array}\right.
\end{array} . \begin{array}{l}
(b)
\end{array}\right.}
\end{aligned}
$$

for any $x \in(\mathrm{a}, \mathrm{b})$
(ii) We also have

$$
\begin{align*}
& \left|\frac{I_{b-, g}^{\alpha} f(a)+I_{a+, g}^{\alpha} f(b)}{2}-\frac{1}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha} \frac{f(b)+f(a)}{2}\right| \\
& \leq \frac{H}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{g^{\prime}(t)(b-t)^{r} d t}{[g(b)-g(t)]^{1-\alpha}}+\int_{a}^{b} \frac{g^{\prime}(t)(t-a)^{r} d t}{[g(t)-g(a)]^{1-\alpha}}\right]  \tag{27}\\
& \leq \frac{H}{\Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}(b-a)^{r}
\end{align*}
$$

Proof. Using the identity (13) and the properties of the modulus, we have

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f(x)+I_{b-, g}^{\alpha} f(x)-\frac{[g(x)-g(a)]^{\alpha} f(a)+[g(b)-g(x)]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)|f(t)-f(a)| d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)|f(t)-f(b)| d t}{[g(t)-g(x)]^{1-\alpha}}\right]=: C(x)
\end{aligned}
$$

for any $x \in(a, b)$.
Since $f:[a, b] \rightarrow \mathbb{C}$ is $r$-H-Hölder continuous on $[a, b]$ with $r \in(0,1]$ and $H>0$, hence
$C(x) \leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{g^{\prime}(t)(t-a)^{r} d t}{[g(x)-g(t)]^{1-\alpha}}+\int_{x}^{b} \frac{g^{\prime}(t)(b-t)^{r} d t}{[g(t)-g(x)]^{1-\alpha}}\right]$

$$
\begin{aligned}
& \leq \frac{H}{\Gamma(\alpha)}\left[(x-a)^{r} \int_{a}^{x} \frac{g^{\prime}(t) d t}{[g(x)-g(t)]^{1-\alpha}}+(b-x)^{r} \int_{x}^{b} \frac{g^{\prime}(t) d t}{[g(t)-g(x)]^{1-\alpha}}\right] \\
& =\frac{H}{\Gamma(\alpha)}\left[(x-a)^{r} \frac{(g(x)-g(a))^{\alpha}}{\alpha}+(b-x)^{r} \frac{(g(b)-g(x))^{\alpha}}{\alpha}\right] \\
& =\frac{H}{\Gamma(\alpha+1)}\left[(x-a)^{r}(g(x)-g(a))^{\alpha}+(b-x)^{r}(g(b)-g(x))^{\alpha}\right],
\end{aligned}
$$

for any $x \in(a, b)$, which proves the first two inequalities in (25). The rest is obvious.

The inequality (26) follows in a similar way by utilising the equality (14).
The inequality (27) follows by utilising the equality (15).
Corollary 7 With the assumptions of Theorem 4 we have

$$
\begin{aligned}
& \left|I_{a+, g}^{\alpha} f\left(M_{g}(a, b)\right)+I_{b-, g}^{\alpha} f\left(M_{g}(a, b)\right)-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t)(t-a)^{r} d t}{\left[g\left(M_{g}(a, b)\right)-g(t)\right]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t)(b-t)^{r} d t}{\left[g(t)-g\left(M_{g}(a, b)\right)\right]^{1-\alpha}}\right] \\
& \leq \frac{H}{2^{\alpha} \Gamma(\alpha+1)}(g(b)-g(a))^{\alpha}\left[\left(M_{g}(a, b)-a\right)^{r}+\left(b-M_{g}(a, b)\right)^{r}\right]
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|I_{M_{g}(a, b)-, g}^{\alpha} f(a)+I_{M_{g}(a, b)+, g}^{\alpha} f(b)-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}[g(b)-g(a)]^{\alpha}\right| \\
& \leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{M_{g}(a, b)} \frac{g^{\prime}(t)(t-a)^{r} d t}{[g(t)-g(a)]^{1-\alpha}}+\int_{M_{g}(a, b)}^{b} \frac{g^{\prime}(t)(b-t)^{r} d t}{[g(b)-g(t)]^{1-\alpha}}\right] \\
& \leq \frac{H}{2^{\alpha} \Gamma(\alpha+1)}(g(b)-g(a))^{\alpha}\left[\left(M_{g}(a, b)-a\right)^{r}+\left(b-M_{g}(a, b)\right)^{r}\right] .
\end{aligned}
$$

## 6 Applications for Hadamard fractional integrals

If we take $\mathrm{g}(\mathrm{t})=\ln \mathrm{t}$ and $0 \leq \mathrm{a}<\mathrm{x} \leq \mathrm{b}$, then by Theorem 3 for Hadamard fractional integrals $H_{a+}^{\alpha}$ and $H_{b-}^{\alpha}$ we have for $f:[a, b] \rightarrow \mathbb{C}$, a function of bounded variation on $[\mathrm{a}, \mathrm{b}]$ that

$$
\left|H_{a+}^{\alpha} f(x)+H_{b-f}^{\alpha} f(x)-\frac{\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha} f(a)+\left[\ln \left(\frac{b}{x}\right)\right]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right|
$$

$$
\left.\begin{array}{l}
\leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{\left[\ln \left(\frac{x}{t}\right)\right]^{\alpha-1} \bigvee_{a}^{t}(f) d t}{t}+\int_{x}^{b} \frac{\left[\ln \left(\frac{t}{x}\right)\right]^{\alpha-1} \bigvee_{t}^{b}(f) d t}{t}\right] \\
\leq \frac{1}{\Gamma(\alpha+1)}\left[\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha} \bigvee_{a}^{x}(f)+\left[\ln \left(\frac{b}{x}\right)\right]_{x}^{\alpha} \bigvee_{x}^{b}(f)\right]
\end{array}\right] \begin{aligned}
& \leq \frac{1}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
{\left[\frac{1}{2} \ln \left(\frac{b}{a}\right)+\left|\ln \left(\frac{x}{G(a, b)}\right)\right|\right]^{\alpha} \bigvee_{a}^{b}(f) ;} \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha p}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha p}\right)^{1 / p}\left(\left(\bigvee_{a}^{x}(f)\right)^{q}+\left(\bigvee_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha}\right)\left[\frac{1}{2} \bigvee_{a}^{b}(f)+\frac{1}{2}\left|\bigvee_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{array}\right.
\end{aligned}
$$

and

$$
\left.\begin{array}{l}
\left|H_{x-}^{\alpha} f(a)+H_{x+}^{\alpha} f(b)-\frac{\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha} f(a)+\left[\ln \left(\frac{b}{x}\right)\right]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
\leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{\left[\ln \left(\frac{t}{a}\right)\right]^{\alpha-1} V_{a}^{t}(f) d t}{t}+\int_{x}^{b} \frac{\left[\ln \left(\frac{b}{t}\right)\right]^{\alpha-1} V_{t}^{b}(f) d t}{t}\right] \\
\leq \frac{1}{\Gamma(\alpha+1)}\left[\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha} \bigvee_{a}^{x}(f)+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha} \bigvee_{x}^{b}(f)\right]
\end{array}\right\} \begin{aligned}
& \leq \frac{1}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
{\left[\frac{1}{2} \ln \left(\frac{b}{a}\right)+\left|\ln \left(\frac{x}{G(a, b)}\right)\right|\right]^{\alpha} \bigvee_{a}^{b}(f) ;} \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha p}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha p}\right)^{1 / p}\left(\left(\bigvee_{a}^{x}(f)\right)^{q}+\left(V_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha}\right)\left[\frac{1}{2} \bigvee_{a}^{b}(f)+\frac{1}{2}\left|V_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{array}\right.
\end{aligned}
$$

for any $x \in(a, b)$
We also have

$$
\begin{aligned}
& \left|\frac{H_{b-}^{\alpha} f(a)+H_{a+}^{\alpha} f(b)}{2}-\frac{1}{\Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha} \frac{f(b)+f(a)}{2}\right| \\
& \leq \frac{1}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{\left[\ln \left(\frac{b}{t}\right)\right]^{\alpha-1} \bigvee_{t}^{b}(f) d t}{t}+\int_{a}^{b} \frac{\left[\ln \left(\frac{t}{a}\right)\right]^{\alpha-1} g^{\prime}(t) \bigvee_{a}^{t}(f) d t}{t}\right] \\
& \leq \frac{1}{\Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha} \bigvee_{a}^{b}(f) .
\end{aligned}
$$

If we take in (28) and (29) $x=G(a, b)$, then we get

$$
\begin{aligned}
& \left|H_{a+}^{\alpha} f(G(a, b))+H_{b-}^{\alpha} f(G(a, b))-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{G(a, b)} \frac{\left[\ln \left(\frac{G(a, b)}{t}\right)\right]^{\alpha-1} V_{a}^{t}(f) d t}{t}+\int_{G(a, b)}^{b} \frac{\left[\ln \left(\frac{t}{G(a, b)}\right)\right]^{\alpha-1} V_{t}^{b}(f) d t}{t}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha} \bigvee_{a}^{b}(f)
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|H_{G(a, b)-}^{\alpha} f(a)+H_{G(a, b)+}^{\alpha} f(b)-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha}\right| \\
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{G(a, b)} \frac{\left[\ln \left(\frac{t}{a}\right)\right]^{\alpha-1} \bigvee_{a}^{t}(f) d t}{t}+\int_{G(a, b)}^{b} \frac{\left[\ln \left(\frac{b}{t}\right)\right]^{\alpha-1} \bigvee_{t}^{b}(f) d t}{t}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]_{a}^{\alpha} \bigvee_{a}^{b}(f) .
\end{aligned}
$$

Assume that $f:[a, b] \rightarrow \mathbb{C}$ is $r$ - H -Hölder continuous on $[\mathrm{a}, \mathrm{b}]$ with $\mathrm{r} \in(0,1]$ and $\mathrm{H}>0$. If we take $\mathrm{g}(\mathrm{t})=\ln \mathrm{t}$ and $0 \leq \mathrm{a}<\mathrm{x} \leq \mathrm{b}$ in Theorem 4, then we get

$$
\begin{aligned}
& \left|H_{a+}^{\alpha} f(x)+H_{b-}^{\alpha} f(x)-\frac{\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha} f(a)+\left[\ln \left(\frac{b}{x}\right)\right]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
& \leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{\left[\ln \left(\frac{x}{t}\right)\right]^{\alpha-1}(t-a)^{r} d t}{t}+\int_{x}^{b} \frac{\left[\ln \left(\frac{t}{x}\right)\right]^{\alpha-1}(b-t)^{r} d t}{t}\right] \\
& \leq \frac{H}{\Gamma(\alpha+1)}\left[\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha}(x-a)^{r}+\left[\ln \left(\frac{b}{x}\right)\right]^{\alpha}(b-x)^{r}\right] \\
& \leq \frac{H}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha p}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha p}\right)^{1 / p}\left(\left(V_{a}^{x}(f)\right)^{q}+\left(V_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
w i t h p, q>1, \frac{1}{p}+\frac{1}{q}=1 ; \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha}\right)\left[\frac{1}{2} V_{a}^{b}(f)+\frac{1}{2}\left|V_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{array}\right.
\end{aligned}
$$

and

$$
\left.\begin{array}{l}
\left|H_{x-}^{\alpha} f(a)+H_{x+}^{\alpha} f(b)-\frac{\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha} f(a)+\left[\ln \left(\frac{b}{x}\right)\right]^{\alpha} f(b)}{\Gamma(\alpha+1)}\right| \\
\leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{x} \frac{\left[\ln \left(\frac{t}{a}\right)\right]^{\alpha-1}(t-a)^{r} d t}{t}+\int_{x}^{b} \frac{\left[\ln \left(\frac{b}{t}\right)\right]^{\alpha-1}(b-t)^{r} d t}{t}\right] \\
\leq \frac{H}{\Gamma(\alpha+1)}\left[\left[\ln \left(\frac{x}{a}\right)\right]^{\alpha}(x-a)^{r}+\left[\ln \left(\frac{b}{x}\right)\right]^{\alpha}(b-x)^{r}\right]
\end{array}\right\} \begin{aligned}
& \leq \frac{H}{\Gamma(\alpha+1)}\left\{\begin{array}{l}
{\left[\frac{1}{2} \ln \left(\frac{b}{a}\right)+\left|\ln \left(\frac{x}{G(a, b)}\right)\right|\right]^{\alpha} V_{a}^{b}(f) ;} \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha p}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha p}\right)^{1 / p}\left(\left(V_{a}^{x}(f)\right)^{q}+\left(V_{x}^{b}(f)\right)^{q}\right)^{1 / q} \\
\left(\left(\ln \left(\frac{x}{a}\right)\right)^{\alpha}+\left(\ln \left(\frac{b}{x}\right)\right)^{\alpha}\right)\left[\frac{1}{2} \bigvee_{a}^{b}(f)+\frac{1}{2}\left|V_{a}^{x}(f)-V_{x}^{b}(f)\right|\right]
\end{array}\right.
\end{aligned}
$$

for any $x \in(a, b)$.
We also have

$$
\begin{align*}
& \left|\frac{H_{b--}^{\alpha} f(a)+H_{a+}^{\alpha} f(b)}{2}-\frac{1}{\Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha} \frac{f(b)+f(a)}{2}\right| \\
& \leq \frac{H}{2 \Gamma(\alpha)}\left[\int_{a}^{b} \frac{\left[\ln \left(\frac{b}{t}\right)\right]^{\alpha-1}(b-t)^{r} d t}{t}+\int_{a}^{b} \frac{\left[\ln \left(\frac{t}{a}\right)\right]^{\alpha-1}(t-a)^{r} d t}{t}\right]  \tag{31}\\
& \leq \frac{H}{\Gamma(\alpha+1)}(b-a)^{r}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha} .
\end{align*}
$$

If we take in (30) and (31) $x=G(a, b)$, then we get

$$
\begin{aligned}
& \left|H_{a+}^{\alpha} f(G(a, b))+H_{b-}^{\alpha} f(G(a, b))-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha}\right| \\
& \leq \frac{H}{\Gamma(\alpha)}\left[\int_{a}^{G(a, b)} \frac{\left[\ln \left(\frac{G(a, b)}{t}\right)\right]^{\alpha-1}(t-a)^{r} d t}{t}+\int_{G(a, b)}^{b} \frac{\left[\ln \left(\frac{t}{G(a, b)}\right)\right]^{\alpha-1}(b-t)^{r} d t}{t}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha}(b-a)^{r}
\end{aligned}
$$

and

$$
\left|H_{G(a, b)-}^{\alpha} f(a)+H_{G(a, b)+}^{\alpha} f(b)-\frac{f(a)+f(b)}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha}\right|
$$

$$
\begin{aligned}
& \leq \frac{1}{\Gamma(\alpha)}\left[\int_{a}^{G(a, b)} \frac{\left[\ln \left(\frac{t}{a}\right)\right]^{\alpha-1}(t-a)^{r} d t}{t}+\int_{G(a, b)}^{b} \frac{\left[\ln \left(\frac{b}{t}\right)\right]^{\alpha-1}(b-t)^{r} d t}{t}\right] \\
& \leq \frac{1}{2^{\alpha} \Gamma(\alpha+1)}\left[\ln \left(\frac{b}{a}\right)\right]^{\alpha}(b-a)^{r} .
\end{aligned}
$$

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## References

[1] A. Aglić Aljinović, Montgomery identity and Ostrowski type inequalities for Riemann-Liouville fractional integral, J. Math., 2014, Art. ID 503195, 6 pp.
[2] A. O. Akdemir, Inequalities of Ostrowski's type for $m$ - and ( $\alpha, \mathfrak{m}$ )logarithmically convex functions via Riemann-Liouville fractional integrals, J. Comput. Anal. Appl., 16 (2014), no. 2, 375-383
[3] G. A. Anastassiou, Fractional representation formulae under initial conditions and fractional Ostrowski type inequalities, Demonstr. Math., 48 (2015), no. 3, 357-378
[4] G. A. Anastassiou, The reduction method in fractional calculus and fractional Ostrowski type inequalities,Indian J. Math., 56 (2014), no. 3, 333357.
[5] H. Budak, M. Z. Sarikaya, E. Set, Generalized Ostrowski type inequalities for functions whose local fractional derivatives are generalized s-convex in the second sense, J. Appl. Math. Comput. Mech., 15 (2016), no. 4, 11-21.
[6] P. Cerone and S. S. Dragomir, Midpoint-type rules from an inequalities point of view, Handbook of analytic-computational methods in applied mathematics, 135-200, Chapman \& Hall/CRC, Boca Raton, FL, 2000.
[7] S. S. Dragomir, The Ostrowski's integral inequality for Lipschitzian mappings and applications, Comput. Math. Appl., 38 (1999), no. 11-12, 33-37.
[8] S. S. Dragomir, The Ostrowski integral inequality for mappings of bounded variation, Bull. Austral. Math. Soc. 60 (1999), No. 3, 495-508.
[9] S. S. Dragomir, On the midpoint quadrature formula for mappings with bounded variation and applications, Kragujevac J. Math., 22 (2000), 1319.
[10] S. S. Dragomir, On the Ostrowski's integral inequality for mappings with bounded variation and applications, Math. Ineq. Appl., 4 (2001), No. 1, 59-66. Preprint: RGMIA Res. Rep. Coll. 2 (1999), Art. 7, [Online: http://rgmia.org/papers/v2n1/v2n1-7.pdf]
[11] S. S. Dragomir, Refinements of the Ostrowski inequality in terms of the cumulative variation and applications, Analysis (Berlin) 34 (2014), No. 2, 223-240. Preprint: RGMIA Res. Rep. Coll., 16 (2013), Art. 29 [Online:http://rgmia.org/papers/v16/v16a29.pdf].
[12] S. S. Dragomir, Ostrowski type inequalities for Lebesgue integral: a survey of recent results, Australian J. Math. Anal. Appl., Volume 14, Issue 1, Article 1, pp. 1-287, 2017. [Online http://ajmaa.org/cgi-bin/paper.pl?string=v14n1/V14I1P1.tex].
[13] S. S. Dragomir, Ostrowski type inequalities for Riemann-Liouville fractional integrals of bounded variation, Hölder and Lipschitzian functions, Preprint RGMIA Res. Rep. Coll., 20 (2017), Art. 48. [Online http://rgmia.org/papers/v20/v20a48.pdf].
[14] S. S. Dragomir, Ostrowski type inequalities for generalized RiemannLiouville fractional integrals of functions with bounded variation, Preprint RGMIA Res. Rep. Coll., 20 (2017), Art 58. [Online http://rgmia.org/papers/v20/v20a58.pdf].
[15] S. S. Dragomir, M. S. Moslehian and Y. J. Cho, Some reverses of the Cauchy-Schwarz inequality for complex functions of self-adjoint operators in Hilbert spaces,Math. Inequal. Appl., 17 (2014), no. 4, 1365-1373.
[16] A. Guezane-Lakoud and F. Aissaoui, New fractional inequalities of Ostrowski type, Transylv. J. Math. Mech., 5 (2013), no. 2, 103-106
[17] A. Kashuri and R. Liko, Ostrowski type fractional integral inequalities for generalized ( $s, m, \varphi$ )-preinvex functions, Aust. J. Math. Anal. Appl., 13 (2016), no. 1, Art. 16, 11 pp.
[18] A. Kilbas, A; H. M. Srivastava and J. J. Trujillo, Theory and Applications of Fractional Differential Equations, North-Holland Mathematics Studies,
204. Elsevier Science B.V., Amsterdam, 2006. xvi+523 pp. ISBN: 978-0-444-51832-3; 0-444-51832-0
[19] M. A. Noor, K. I.Noor and S. Iftikhar, Fractional Ostrowski inequalities for harmonic h-preinvex functions, Facta Univ. Ser. Math. Inform., 31 (2016), no. 2, 417-445
[20] M. Z. Sarikaya and H. Filiz, Note on the Ostrowski type inequalities for fractional integrals, Vietnam J. Math., 42 (2014), no. 2, 187-190
[21] M. Z. Sarikaya and H. Budak, Generalized Ostrowski type inequalities for local fractional integrals, Proc. Amer. Math. Soc., 145 (2017), no. 4, 1527-1538.
[22] E. Set, New inequalities of Ostrowski type for mappings whose derivatives are s-convex in the second sense via fractional integrals, Comput. Math. Appl., 63 (2012), no. 7, 1147-1154.
[23] M. Tunç, On new inequalities for h-convex functions via RiemannLiouville fractional integration, Filomat, 27:4 (2013), 559-565.
[24] M. Tunç, Ostrowski type inequalities for $m$ - and ( $\alpha, m$ )-geometrically convex functions via Riemann-Louville fractional integrals, Afr. Mat., 27 (2016), no. 5-6, 841-850.
[25] H. Yildirim and Z. Kirtay, Ostrowski inequality for generalized fractional integral and related inequalities, Malaya J. Mat., 2 (3) (2014), 322-329.
[26] C. Yildiz, E, Özdemir and Z. S. Muhamet, New generalizations of Ostrowski-like type inequalities for fractional integrals, Kyungpook Math. J. 56 (2016), no. 1, 161-172.
[27] H. Yue, Ostrowski inequality for fractional integrals and related fractional inequalities, Transylv. J. Math. Mech., 5 (2013), no. 1, 85-89.

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# Composition iterates, Cauchy, translation, and Sincov inclusions 

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#### Abstract

Improving and extending some ideas of Gottlob Frege from 1874 (on a generalization of the notion of the composition iterates of a function), we consider the composition iterates $\varphi^{n}$ of a relation $\varphi$ on $X$, defined by


$$
\varphi^{0}=\Delta_{X}, \quad \varphi^{n}=\varphi \circ \varphi^{n-1} \quad \text { if } \quad n \in \mathbb{N}, \quad \text { and } \quad \varphi^{\infty}=\bigcup_{n=0}^{\infty} \varphi^{n}
$$

In particular, by using the relational inclusion $\varphi^{n} \circ \varphi^{m} \subseteq \varphi^{n+m}$ with $n, m \in \overline{\mathbb{N}}_{0}=\{0\} \cup \mathbb{N} \cup\{\infty\}$, we show that the function $\alpha$, defined by

$$
\alpha(n)=\varphi^{n} \quad \text { for } \quad n \in \overline{\mathbb{N}}_{0}
$$

satisfies the Cauchy problem

$$
\alpha(n) \circ \alpha(m) \subseteq \alpha(n+m), \quad \alpha(0)=\Delta_{x}
$$

Moreover, the function $f$, defined by

$$
f(n, A)=\alpha(n)[A] \quad \text { for } \quad n \in \overline{\mathbb{N}}_{0} \quad \text { and } \quad A \subseteq X
$$

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satisfies the translation problem

$$
f(n, f(m, A)) \subseteq f(n+m, A), \quad f(0, A)=A
$$

Furthermore, the function F, defined by

$$
F(A, B)=\left\{n \in \overline{\mathbb{N}}_{0}: \quad A \subseteq f(n, B)\right\} \quad \text { for } \quad A, B \subseteq X
$$

satisfies the Sincov problem

$$
F(A, B)+F(B, C) \subseteq F(A, C), \quad 0 \in F(A, A)
$$

Motivated by the above observations, we investigate a function $F$ on the product set $\mathrm{X}^{2}$ to the power groupoid $\mathcal{P}(\mathrm{U})$ of an additively written groupoid U which is supertriangular in the sense that

$$
F(x, y)+F(y, z) \subseteq F(x, z)
$$

for all $x, y, z \in X$. For this, we introduce the convenient notations

$$
R(x, y)=F(y, x) \quad \text { and } \quad S(x, y)=F(x, y)+R(x, y)
$$

and

$$
\Phi(x)=F(x, x) \quad \text { and } \quad \Psi(x)=\bigcup_{y \in X} S(x, y)
$$

Moreover, we gradually assume that $U$ and $F$ have some useful additional properties. For instance, U has a zero, U is a group, U is commutative, U is cancellative, or U has a suitable distance function; while F is nonpartial, $F$ is symmetric, skew symmetric, or single-valued.

## 1 A few basic facts on relations

In [40], a subset $F$ of a product set $X \times Y$ is called a relation on $X$ to $Y$. In particular, a relation on $X$ to itself is called a relation on $X$. More specially, $\Delta_{X}=\{(x, x): x \in X\}$ is called the identity relation on $X$.

If $F$ is a relation on $X$ to $Y$, then by the above definitions we can also state that $F$ is a relation on $X \cup Y$. However, for our present purposes, the latter view of the relation $F$ would also be quite unnatural.

If $F$ is a relation on $X$ to $Y$, then for any $x \in X$ and $A \subseteq X$ the sets $F(x)=\{y \in Y:(x, y) \in F\}$ and $F[A]=\bigcup_{a \in A} F(a)$ are called the images of $x$ and $\mathcal{A}$ under $F$, respectively.

If $(x, y) \in F$, then instead of $y \in F(x)$, we may also write $x F y$. However, instead of $F[A]$, we cannot write $F(A)$. Namely, it may occur that, in addition to $A \subseteq X$, we also have $A \in X$.

The sets $D_{F}=\{x \in X: F(x) \neq \emptyset\}$ and $R_{F}=F[X]$ are called the domain and range of F , respectively. If in particular $\mathrm{D}_{\mathrm{F}}=\mathrm{X}$, then we say that F is a relation of X to Y , or that F is a nonpartial relation on X to Y .

In particular, a relation $f$ on $X$ to $Y$ is called a function if for each $X \in D_{f}$ there exists $y \in Y$ such that $f(x)=\{y\}$. In this case, by identifying singletons with their elements, we may simply write $f(x)=y$ instead of $f(x)=\{y\}$.

In particular, a function $\star$ of $X$ to itself is called a unary operation on $X$, while a function $*$ of $X^{2}$ to $X$ is called a binary operation on $X$. In this case, for any $x, y \in X$, we usually write $x^{\star}$ and $x * y$ instead of $\star(x)$ and $*((x, y))$.

If $F$ is a relation on $X$ to $Y$, then we can easily see that $F=\bigcup_{x \in X}\{x\} \times F(x)$. Therefore, the values $F(x)$, where $x \in X$, uniquely determine $F$. Thus, a relation $F$ on $X$ to $Y$ can also be naturally defined by specifying $F(x)$ for all $x \in X$.

For instance, the inverse $\mathrm{F}^{-1}$ can be defined such that $\mathrm{F}^{-1}(\mathrm{y})=\{x \in X: y \in$ $F(x)\}$ for all $y \in Y$. Moreover, if $G$ is a relation on $Y$ to $Z$, then the composition $G \circ F$ can be defined such that $(G \circ F)(x)=G[F(x)]$ for all $x \in X$.
If $F$ is a relation on $X$ to $Y$, then a relation $\Phi$ of $D_{F}$ to $Y$ is called a selection relation of F if $\Phi \subseteq \mathrm{F}$, i.e., $\Phi(x) \subseteq \mathrm{F}(\mathrm{x})$ for all $\mathrm{x} \in \mathrm{D}_{\mathrm{F}}$. By using the Axiom of Choice, it can be seen that every relation is the union of its selection functions.

For a relation F on X to Y , we may naturally define two set-valued functions $\varphi$ of $X$ to $\mathcal{P}(Y)$ and $\Phi$ of $\mathcal{P}(X)$ to $\mathcal{P}(Y)$ such that $\varphi(x)=F(x)$ for all $x \in X$ and $\Phi(A)=F[A]$ for all $A \subseteq X$.

Functions of $X$ to $\mathcal{P}(Y)$ can be identified with relations on $X$ to $Y$, while functions of $\mathcal{P}(\mathrm{X})$ to $\mathcal{P}(\mathrm{Y})$ are more powerful objects than relations on X to Y . They were briefly called co-relations on X to Y in [40].

In particular, a relation R on X can be briefly defined to be reflexive if $\Delta_{\mathrm{X}} \subseteq R$, and transitive if $R \circ R \subseteq R$. Moreover, $R$ can be briefly defined to be symmetric if $R^{-1} \subseteq R$, and antisymmetric if $R \cap R^{-1} \subseteq \Delta_{x}$.

Thus, a reflexive and transitive (symmetric) relation may be called a preorder (tolerance) relation, and a symmetric (antisymmetric) preorder relation may be called an equivalence (partial order) relation.

For $A \subseteq X$, Pervin's relation $R_{A}=A^{2} \cup A^{c} \times X$, with $A^{c}=X \backslash A$, is an important preorder on X . While, for a pseudometric d on X , Weil's surrounding $B_{r}=\left\{(x, y) \in X^{2}: d(x, y)<r\right\}$, with $r>0$, is an important tolerance on $X$.

Note that $S_{A}=R_{A} \cap R_{A}^{-1}=R_{A} \cap R_{A^{c}}=A^{2} \cap\left(A^{c}\right)^{2}$ is already an equivalence on $X$. And, more generally if $\mathcal{A}$ is a partition of $X$, then $S_{\mathcal{A}}=\bigcup_{\mathcal{A} \in \mathcal{A}} A^{2}$ is an equivalence on X which can, to some extent, be identified with $\mathcal{A}$.

## 2 A few basic facts on ordered sets and groupoids

If $\leq$ is a relation on $X$, then motivated by Birkhoff [5, p. 1] the ordered pair $X(\leq)=(X, \leq)$ is called a goset (generalized ordered set) [39]. In particular, it is called a proset (preordered set) if the relation $\leq$ is a preorder on X .

Quite similarly, a goset $X(\leq)$ is called a poset (partially ordered set) if the relation $\leq$ is a partial order on X . The importance of posets lies mainly in the fact that any family of sets forms a poset with set inclusion.

A function $f$ of one goset $\mathrm{X}(\leq)$ to another $\mathrm{Y}(\leq)$ is called increasing if $\mathrm{x}_{1} \leq \mathrm{x}_{2}$ implies $f\left(x_{1}\right) \leq f\left(x_{2}\right)$ for all $x_{1}, x_{2} \in X$. The function $f$ can now be briefly called decreasing if it is increasing as a function of $X(\leq)$ to the dual $Y(\geq)$.

An increasing function $\varphi$ of the goset $X=X(\leq)$ to itself is called a projection (involution) operation on X if it is idempotent (involutive) in the sense that $\varphi \circ \varphi=\varphi\left(\varphi \circ \varphi=\Delta_{X}\right)$. Note that $\varphi \circ \varphi=\Delta_{X}$ if and only if $\varphi^{-1}=\varphi$.

Moreover, a projection operation $\varphi$ on a poset $X$ is called a closure operation on X if it is extensive in the sense that $\Delta_{\mathrm{X}} \leq \varphi$. That is, $\mathrm{x} \leq \varphi(\mathrm{x})$ for all $x \in \mathrm{X}$. The interior operations can again be most briefly defined by dualization.

If $f$ is a function of one goset $X$ to another $Y$ and $g$ is a function of $Y$ to $X$ such that, for any $x \in X$ and $y \in Y$, we have $f(x) \leq y$ if and only $x \leq g(y)$, then g is called a Galois adjoint of $\mathrm{f}[12, \mathrm{p} .155]$.

Hence, by taking $\varphi=\mathrm{g} \circ \mathrm{f}$, one can easily see that, for any $\mathfrak{u}, v \in X$, we have $f(u) \leq f(v)$ if and only if $u \leq \varphi(v)$. Moreover, if $X$ and $Y$ are prosets, then it can be shown that $f$ is increasing, $\varphi$ is a closure and $f=f \circ \varphi[39]$.

If + is a binary operation on a set $X$, then the ordered pair $X(+)=(X,+)$ is called an additive groupoid. Recently, groupoids are usually called magmas, not to be confused with Brandt groupoids [6].

If $X$ is a groupoid, then for any $A, B \subseteq X$ we may also naturally define $A+B=\{x+y: x \in A, y \in B\}$. Thus, by identifying singletons with their elements, X may be considered as a subgoupoid of its power groupoid $\mathcal{P}(\mathrm{X})$.

In a groupoid $X$, for any $n \in \mathbb{N}$ and $x \in X$ we may also naturally define $n x=x$ if $n=1$, and $n x=(n-1) x+x$ if $n>1$. Thus, for any $n \in \mathbb{N}$ and $A \subseteq X$, we may also naturally define $n \mathcal{A}=\{n x: x \in A\}$.

If X is a semigroup (associative groupoid), then we have $(\mathrm{n}+\mathfrak{m}) \mathrm{x}=\mathrm{n} \boldsymbol{x}+\mathfrak{m} \boldsymbol{x}$ and $(n m) x=\mathfrak{n}(m x)$ for all $n, m \in \mathbb{N}$ and $x \in X$. However, the equality $\mathfrak{n}(x+y)=n x+n y$ requires the elements $x, y \in X$ to be commuting [19].

If the groupoid $X$ has a zero element 0 , then we also naturally define $0 x=0$ for all $x \in X$. Moreover, if $X$ is a group, then we also naturally define $(-\mathfrak{n}) x=$ $n(-x)$ for all $n \in \mathbb{N}$ and $x \in X$. And thus also $k A$ for all $k \in \mathbb{Z}$ and $A \subseteq X$.

Concerning the corresponding operations in $\mathcal{P}(X)$, we must be very careful.

Namely, in general, we only have $(n+m) A \subseteq n A+m A$ and $n A \subseteq \sum_{i=1}^{n} A$ for all $n, m \in \mathbb{N}$ and $A \subseteq X$. However, $\mathcal{P}(X)$ has a richer structure than $X$.

In particular, an element $x$ of a groupoid $X$ is called left-cancellable if $x+y=$ $x+z$ implies $y=z$ for all $y, z \in X$. Moreover, the groupoid $X$ is called leftcancellative if every element of $X$ is left-cancellable.
"Right-cancellable" and "right-cancellative" are to be defined quite similarly. Moreover, for instance, the groupoid X is to be called cancellative if it is both left-cancellative and right-cancellative.

A semigroup X can be easily embedded in a monoid (semigroup with zero element), by adjoining an element 0 not in $X$, and defining $0+x=x+0=x$ for all $x \in X$. Important monoids will be $\mathbb{N}_{0}=\{0\} \cup \mathbb{N}$ and $\overline{\mathbb{N}}_{0}=\mathbb{N}_{0} \cup\{\infty\}$.

## 3 The finite composition iterates of a relation

Notation 1 In the sequel, we shall assume that X is a set, $\Delta$ is the identity function of X and $\varphi$ is a relation on X .

Note that the family $\mathcal{P}\left(\mathrm{X}^{2}\right)$ of all relations on X forms a semigroup, with identity element $\Delta$, with respect to the composition of relations. Therefore, we may naturally use the following

Definition 1 Define $\varphi^{0}=\Delta$, and for any $n \in \mathbb{N}$

$$
\varphi^{n}=\varphi \circ \varphi^{n-1} .
$$

Remark 1 Thus, for each $n \in \mathbb{N}_{0}, \varphi^{n}$ is also a relation on $X$ which is called the $n$th composition iterate of $\varphi$.

Now, as a particular case of a more general theorem on monoids, we can state the following theorem whose direct proof is included here only for the reader's convenience.

Theorem 1 For any $\mathrm{n}, \mathrm{m} \in \mathbb{N}_{0}$, we have

$$
\varphi^{\mathrm{n}+\mathrm{m}}=\varphi^{\mathrm{n}} \circ \varphi^{\mathrm{m}} .
$$

Proof. For fixed $\mathfrak{m} \in \mathbb{N}_{0}$, we shall prove, by induction, that

$$
\varphi^{m+n}=\varphi^{n} \circ \varphi^{m}
$$

for all $\mathfrak{n} \in \mathbb{N}_{0}$. Hence, by the commutativity of the addition in $\mathbb{N}_{0}$, the assertion of the theorem follows.

By Definition 1, we evidently have $\varphi^{m+0}=\varphi^{m}=\Delta \circ \varphi^{m}=\varphi^{0} \circ \varphi^{m}$. Therefore, the required equity is true for $n=0$.

Let us suppose now that the required equality is true for some $n \in \mathbb{N}_{0}$. Then, by Definition 1, the above assumption, and the corresponding associativities, we have

$$
\begin{aligned}
\varphi^{\mathfrak{m}+(\mathfrak{n}+1)} & =\varphi^{(\mathfrak{m}+\mathfrak{n})+1}=\varphi \circ \varphi^{\mathfrak{m}+\mathfrak{n}} \\
& =\varphi \circ\left(\varphi^{\mathfrak{n}} \circ \varphi^{\mathfrak{m}}\right)=\left(\varphi \circ \varphi^{\mathfrak{n}}\right) \circ \varphi^{\mathrm{m}}=\varphi^{\mathfrak{n}+1} \circ \varphi^{\mathrm{m}} .
\end{aligned}
$$

Therefore, the required equality is also true for $n+1$.
Remark 2 This theorem shows that the family $\left\{\varphi^{n}\right\}_{n=0}^{\infty}$ also forms a semigroup, with identity element $\Delta$, with respect to composition.

By induction, we can also easily prove the less trivial part of the following
Theorem 2 The following assertions are equivalent:
(1) $\Delta \subseteq \varphi$;
(2) $\varphi^{\mathrm{n}} \subseteq \varphi^{\mathfrak{n}+1}$ for all $\mathrm{n} \in \mathbb{N}_{0}$.

Remark 3 This theorem shows that the sequence $\left(\varphi^{n}\right)_{n=0}^{\infty}$ is increasing, with respect to set inclusion, if and only if the relation $\varphi$ is reflexive on $X$.

Note that if in particular $\varphi$ is reflexive on $X$ and $\varphi$ is a function, then we necessarily have $\varphi=\Delta$, and thus also $\varphi^{n}=\Delta$ for all $n \in \mathbb{N}_{0}$.

Therefore, in the important particular case when $\varphi$ is a function of $X$ to itself, Theorem 2 cannot have any significance.

## 4 The infinite composition iterate of a relation

In addition to Definition 1, we may also naturally use the following
Definition 2 Define

$$
\varphi^{\infty}=\bigcup_{n=0}^{\infty} \varphi^{n} .
$$

Remark 4 Moreover, the relations

$$
\underline{\lim }_{n \rightarrow \infty} \varphi^{n}=\bigcup_{n=0}^{\infty} \bigcap_{k=n}^{\infty} \varphi^{k} \quad \text { and } \quad \varlimsup_{n \rightarrow \infty} \varphi^{n}=\bigcap_{n=0}^{\infty} \bigcup_{k=n}^{\infty} \varphi^{k}
$$

may also be naturally investigated.
Note that if in particular the sequence $\left(\varphi^{n}\right)_{n=0}^{\infty}$ is increasing with respect to set inclusion, then these relations coincide with $\varphi^{\infty}$.

The relation $\varphi^{\infty}$ is called the preorder hull (closure) of $\varphi$. Namely, we have
Theorem $3 \varphi^{\infty}$ is the smallest preorder relation on $X$ containing $\varphi$.
Proof. By Definition 2, it is clear that $\Delta \subseteq \varphi^{\infty}$ and $\varphi \subseteq \varphi^{\infty}$. Thus, $\varphi^{\infty}$ is reflexive and contains $\varphi$.

Moreover, if $(x, y) \in \varphi^{\infty}$ and $(y, z) \in \varphi^{\infty}$, then by Definition 2 there exist $m, n \in \mathbb{N}_{0}$ such that $(x, y) \in \varphi^{m}$ and $(y, z) \in \varphi^{n}$. Hence, by using Theorem 1 , we can infer that $(x, z) \in \varphi^{n} \circ \varphi^{m}=\varphi^{\mathfrak{n}+m}$. Thus, by Definition 2 , we also have $(x, z) \in \varphi^{\infty}$. Therefore, $\varphi^{\infty}$ is also transitive.

On the other hand, if $\psi$ is a relation on $X$ such that $\varphi \subseteq \psi$, then we can note that $\varphi^{n} \subseteq \psi^{n}$ for all $n \in \mathbb{N}_{0}$, and thus by Definition 2 we have $\varphi^{\infty} \subseteq \psi^{\infty}$. Moreover, if $\psi$ is reflexive, then $\psi^{0} \subseteq \psi$. And, if $\psi$ is transitive, then $\psi^{n} \subseteq \psi$ for all $\mathfrak{n} \in \mathbb{N}$. Therefore, if $\psi$ is both reflexive and transitive, then by Definition 2 we have $\psi^{\infty} \subseteq \psi$, and thus also $\varphi^{\infty} \subseteq \psi$.

Now, as an immediate consequence of this theorem, we can also state
Corollary 1 The following assertions are equivalent:
(1) $\varphi^{\infty}=\varphi$;
(2) $\varphi$ is a preorder on X .

Remark 5 From the above results, it is clear that $\infty$ is a closure operation on the poset $\mathcal{P}\left(\mathrm{X}^{2}\right)$.

In general, it is not even finitely union preserving. However, it is compatible with the inversion of relations [18].

Moreover, in addition to Theorem 1, we can also easily prove the following
Theorem 4 For any $n, m \in \overline{\mathbb{N}}_{0}$, we have

$$
\varphi^{n} \circ \varphi^{m} \subseteq \varphi^{n+m}
$$

Moreover, if $\varphi$ is reflexive on X , then the corresponding equality is also true.
Proof. If in particular $n, m \in \mathbb{N}_{0}$, then by Theorem 1 the corresponding equality is true even if $\varphi$ is not assumed to be reflexive.

Moreover, by using Definition 2 and Theorem 3, we can see that

$$
\varphi^{n} \circ \varphi^{\infty} \subseteq \varphi^{\infty} \circ \varphi^{\infty} \subseteq \varphi^{\infty}=\varphi^{n+\infty} .
$$

Furthermore, if $\varphi$ is reflexive, then it is clear that we also have

$$
\varphi^{\infty}=\Delta \circ \varphi^{\infty} \subseteq \varphi^{n} \circ \varphi^{\infty}
$$

Therefore, in this case, $\varphi^{n} \circ \varphi^{\infty}=\varphi^{\infty}=\varphi^{n+\infty}$ also holds. The case " $\infty+m$ " can be treated quite similarly.

Remark 6 Now, in addition to Theorem 2, we can only state that $\varphi^{n} \subseteq \varphi^{\infty}$ for all $\mathfrak{n} \in \overline{\mathbb{N}}_{0}$.

However, by [30], we may naturally say that $\varphi$ is $n$-well-chained if $\varphi^{n}=X^{2}$. And, $\varphi$ is $n$-connected if $\varphi \cup \varphi^{-1}$ is $n$-well-chained.

Moreover, under the notation $\mathcal{T}_{\varphi}=\{A \subseteq X: \varphi[A] \subseteq A\}$ of [24], we have $\varphi^{\infty}=\bigcap_{A \in \mathcal{T}_{\varphi}} R_{A}$. And, $\varphi^{\infty}$ is the largest relation on $X$ such that $\mathcal{T}_{\varphi^{\infty}}=\mathcal{T}_{\varphi}$.

## 5 From the composition iterates to a Cauchy inclusion

Now, extending an idea of Frege $[15,16]$, we may also naturally introduce
Definition 3 For any $\mathfrak{n} \in \overline{\mathbb{N}}_{0}$, define

$$
\alpha(n)=\varphi^{n}
$$

Thus, $\alpha$ may be considered as a relation on $\overline{\mathbb{N}}_{0}$ to $X^{2}$, or as a function of $\overline{\mathbb{N}}_{0}$ to $\mathcal{P}\left(X^{2}\right)$, which can be proved to satisfy a Cauchy type inclusion.

First of all, by Theorem 1, we evidently have the following
Theorem 5 For any $\mathfrak{n}, \mathfrak{m} \in \mathbb{N}_{0}$, we have

$$
\alpha(n+m)=\alpha(n) \circ \alpha(m)
$$

Proof. By Definition 3 and Theorem 1, it is clear that

$$
\alpha(n+m)=\varphi^{n+m}=\varphi^{n} \circ \varphi^{m}=\alpha(n) \circ \alpha(m)
$$

Remark 7 In addition to this theorem, it is also worth noticing that $\alpha(0)=\Delta$.
Moreover, by Theorem 2, we can also at once state the following

Theorem 6 The following assertions are equivalent:
(1) $\Delta \subseteq \varphi$;
(2) $\alpha(n) \subseteq \alpha(n+1)$ for all $n \in \mathbb{N}_{0}$.

Remark 8 Thus, the restriction of the set-valued function $\alpha$ to $\mathbb{N}_{0}$ is increasing, with respect to set inclusion, if and only if the relation $\varphi$ is reflexive on $X$.

By using Theorem 4 instead of Theorem 1, we can also easily establish
Theorem 7 For any $\mathfrak{n}, \mathfrak{m} \in \overline{\mathbb{N}}_{0}$ we have

$$
\alpha(n) \circ \alpha(m) \subseteq \alpha(n+m)
$$

Moreover, if $\varphi$ is reflexive on X , then the corresponding equality is also true.
Remark 9 Now, in addition to Theorem 6, we can also state that $\alpha(n) \subseteq$ $\alpha(\infty)$ for all $n \in \overline{\mathbb{N}}_{0}$.

Thus, in particular, the set-valued function $\alpha$ is increasing, with respect to set inclusion, if and only if the relation $\varphi$ is reflexive on $X$.

## 6 From a Cauchy inclusion to a translation inclusion

Now, as an extension of our former observations, we may naturally start with
Notation 2 Suppose that U is a additive groupoid and $\alpha$ is a relation on U to $X^{2}$ such that

$$
\alpha(u) \circ \alpha(v) \subseteq \alpha(u+v)
$$

for all $\mathbf{u}, \boldsymbol{v} \in \mathbf{U}$.
Thus, extending an idea of Frege $[15,16]$, we may also naturally introduce
Definition 4 For any $u \in U$ and $A \subseteq X$, define

$$
f(u, A)=\alpha(u)[A]
$$

Thus, f may be considered a relation on $\mathrm{U} \times \mathcal{P}(X)$ to $X$, or as a function of $\mathrm{U} \times \mathcal{P}(\mathrm{X})$ to $\mathcal{P}(\mathrm{X})$, which can be proved to satisfy a translation inclusion.

Theorem 8 For any $u, v \in \mathbb{U}$ and $\mathrm{A} \subseteq X$, we have

$$
f(u, f(v, A)) \subseteq f(u+v, A)
$$

Proof. By Definition 4 and the assumed superadditivity property of $\alpha$, we have

$$
\begin{aligned}
& f(u, f(v, A))=\alpha(u)[f(v, A)]=\alpha(u)[\alpha(v)[A]] \\
& \\
& =(\alpha(u) \circ \alpha(v))[A] \subseteq \alpha(u+v)[A]=f(u+v, A) .
\end{aligned}
$$

Remark 10 Thus, by identifying singleton with their elements, we may also write

$$
f(u, f(v, x)) \subseteq f(u+v, x)
$$

for all $u, v \in U$ and $x \in X$.
Now, to illustrate the appropriateness of Definition 4, we can also state
Example 1 If in particular $\alpha$ is as in Definition 3, then by Definition 4 we have

$$
\mathrm{f}(\mathrm{n}, \mathrm{~A})=\alpha(\mathrm{n})[\mathcal{A}]=\varphi^{\mathrm{n}}[\mathrm{~A}]
$$

for all $\mathrm{n} \in \overline{\mathbb{N}}_{0}$ and $\mathrm{A} \subseteq X$. Thus, in particular $\mathrm{f}(0, \mathcal{A})=\mathcal{A}$ for all $\mathrm{A} \subseteq X$.

## 7 From a translation inclusion to a Sincov inclusion

Now, as an extension of our former observations, we may also naturally start with the following

Notation 3 Suppose that U is an additive groupoid, X is a goset and f is a function of $\mathrm{U} \times \mathrm{X}$ to X such that f is increasing in its second variable and

$$
\mathfrak{f}(u, f(v, x)) \leq f(u+v, x)
$$

for all $\mathfrak{u}, v \in \mathrm{U}$ and $x \in \mathrm{X}$.
Thus, improving an idea of Frege [15, 16], we may also naturally introduce
Definition 5 For any $x, y \in X$, define

$$
F(x, y)=\{u \in U: \quad x \leq f(u, y)\} .
$$

Thus, F may be considered as a relation on $\mathrm{X}^{2}$ to U , or as a function of $\mathrm{X}^{2}$ to $\mathcal{P}(\mathrm{U})$, which can be proved to satisfy a Sincov type inclusion.

Theorem 9 For any $x, y, z \in X$, we have

$$
F(x, y)+F(y, z) \subseteq F(x, z) .
$$

Proof. If

$$
u \in F(x, y) \quad \text { and } \quad v \in F(y, z),
$$

then by Definition 5 we get

$$
x \leq f(u, y) \quad \text { and } \quad y \leq f(v, z) .
$$

Hence, by using the assumed increasingness and translation property of $f$, we can infer that

$$
x \leq f(u, y) \leq f(u, f(v, z)) \leq f(u+v, z) .
$$

Therefore, by Definition 5, we have

$$
u+v \in F(x, z) .
$$

Thus, the required inclusion is true.
Now, to illustrate the appropriateness of Definition 5, we can also state
Example 2 If f is as in Example 1, then by Definition 5 we have

$$
F(A, B)=\left\{n \in \overline{\mathbb{N}}_{0}: \quad A \subseteq f(n, B)\right\}=\left\{n \in \overline{\mathbb{N}}_{0}: \quad A \subseteq \varphi^{n}[B]\right\}
$$

for all $\mathrm{A}, \mathrm{B} \subseteq X$. Thus, in particular $0 \in \mathrm{~F}(\mathrm{~A}, \mathrm{~A})$ for all $\mathrm{A} \subseteq \mathrm{X}$.
Remark 11 By Aczél [1, pp. 223, 303 and 353], Sincov's functional equation and its generalizations have been investigated by a surprisingly great number of authors.

For some more recent investigations, see $[4,33,38,27,34,35,7,8,3,14,13]$. The most relevant ones are the set-valued considerations of Smajdor [38] and Augustová and Klapka [3].

Moreover, it is noteworthy that, by using the famous partial operation

$$
(x, y) \bullet(y, z)=(x, z)
$$

the above Sincov inclusion can be turned into a restricted Cauchy inclusion.
Therefore, some of the methods of the theory of superadditive functions and relations $[28,17,29,19]$ can certainly be applied to investigate the corresponding Sincov inequalities and inclusions.

## 8 Some immediate consequences of a Sincov inclusion

Now, motivated by our former observations, we may also naturally introduce the following notations and definitions.

Notation 4 In what follows, we shall also assume that X is a set and U is an additive groupoid. Moreover, we shall suppose that F is a relation on $\mathrm{X}^{2}$ to U .

Definition 6 The relation $F$ will be called supertriangular if

$$
F(x, y)+F(y, z) \subseteq F(x, z)
$$

for all $x, y, z \in X$.
Remark 12 Now, the relation F may also be naturally called subtriangular if the reverse inclusion holds. Moreover, F may be naturally called triangular if it is both subtriangular and supertriangular.

Subtriangular relations are certainly more important than the supertriangular ones. Namely, if a function $d$ of $X^{2}$ to $[0,+\infty]$ satisfies the triangle inequality

$$
\mathrm{d}(\mathrm{x}, \mathrm{z}) \leq \mathrm{d}(\mathrm{x}, \mathrm{y})+\mathrm{d}(\mathrm{y}, z)
$$

for all $x, y, z \in X$, then the relation $F$, defined such that

$$
F(x, y)=[0, d(x, y)] \quad(F(x, y)=[-d(x, y), d(x, y)])
$$

for all $x, y \in X$, can, in general, be proved to be only subtriangular [2].
The $y=x, y=z$ and $z=x$ particular cases of the inclusion considered in Definition 6 strongly suggest the introduction of the following

Definition 7 For any $x, y \in X$, define

$$
R(x, y)=F(y, x) \quad \text { and } \quad S(x, y)=F(x, y)+R(x, y)
$$

Moreover, for any $x \in X$, define

$$
\Phi(x)=F(x, x) \quad \text { and } \quad \Psi(x)=\bigcup_{y \in X} S(x, y)
$$

Thus, R and S may be considered as relations on $\mathrm{X}^{2}$ to U , and $\Phi$ and $\Psi$ may be considered as relations on X to U .

Concerning these relations, we can easily prove the following

Theorem 10 For any $x, y \in X$ we have
(1) $\Phi(x)+\Phi(x) \subseteq \Psi(x)$;
(2) $\mathrm{R}(\mathrm{x}, \mathrm{x})=\Phi(\mathrm{x}) ; \quad$ (3) $\mathrm{S}(\mathrm{x}, \mathrm{x})=\Phi(\mathrm{x})+\Phi(\mathrm{x})$.

Proof. By Definition 7, we evidently have

$$
R(x, x)=F(x, x)=\Phi(x)
$$

and thus also

$$
S(x, x)=F(x, x)+R(x, x)=\Phi(x)+\Phi(x)
$$

Hence, by using the definition of $\Psi$, we can also easily note that

$$
\Phi(x)+\Phi(x)=S(x, x) \subseteq \bigcup_{y \in X} S(x, y)=\Psi(x)
$$

Therefore, assertions (2), (3) and (1) are true.
Now, as a counterpart of [38, Lemma 1] of Wilhelmina Smajdor, we can also prove the following

Theorem 11 If F is supertriangular, then for any $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ we have
(1) $\Psi(x) \subseteq \Phi(x)$;
(2) $\Phi(x)+F(x, y) \subseteq F(x, y)$;
(3) $F(x, y)+\Phi(y) \subseteq F(x, y)$.

Proof. By using Definition 7 and the corresponding particular cases of the inclusion considered in Definition 6, we can easily see that

$$
\Phi(x)+F(x, y)=F(x, x)+F(x, y) \subseteq F(x, y)
$$

and

$$
F(x, y)+\Phi(y)=F(x, y)+F(y, y) \subseteq F(x, y)
$$

Moreover,

$$
S(x, y)=F(x, y)+R(x, y)=F(x, y)+F(y, x) \subseteq F(x, x)=\Phi(x)
$$

and thus also

$$
\Psi(x)=\bigcup_{y \in X} S(x, y) \subseteq \bigcup_{y \in X} \Phi(x) \subseteq \Phi(x)
$$

Therefore, assertions (2), (3) and (1) are true even if only some consequences of the assumed inclusion property of $F$ are supposed to hold.

Now, as an immediate consequence of the above two theorems, we can also state

Corollary 2 If F is supertriangular, then for any $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ we have
(1) $\Phi(x)+\Phi(x) \subseteq \Phi(x)$;
(2) $\Psi(x)+\Psi(x) \subseteq \Psi(x)$;
(3) $\Psi(x)+\Phi(x) \subseteq \Psi(x)$;
(4) $\Phi(x)+\Psi(x) \subseteq \Psi(x)$;
(5) $\Psi(x)+F(x, y) \subseteq F(x, y)$;
(6) $F(x, y)+\Psi(y) \subseteq F(x, y)$.

Remark 13 By [8], in addition to Definition 6, the separability equation

$$
F(x, y)+F(y, z)=F(x, z)+\Phi(y)
$$

may also be naturally investigated.
Moreover, if in particular U is a group, then in addition to Definition 7, the disymmetry relation $D$ of $F$, defined such that $D(x, y)=F(x, y)-R(x, y)$ for all $x, y \in X$, may also be naturally investigated.

## 9 The particular case when $U$ has a zero element

Theorem 12 If F is supertriangular, U has a one-sided zero element 0 and $\mathrm{x} \in \mathrm{X}$ is such that $0 \in \Phi(\mathrm{x})$, then
(1) $\Phi(x)=\Psi(x)$;
(2) $\Phi(x)=\Phi(x)+\Phi(x)$.

Proof. If 0 is a right zero element of U , then by using Theorems 10 and 11 we can see that

$$
\Phi(x)=\Phi(x)+\{0\} \subseteq \Phi(x)+\Phi(x) \subseteq \Psi(x) \subseteq \Phi(x)
$$

While, if 0 is a left zero element of $U$, then we can quite similarly see that

$$
\Phi(x)=\{0\}+\Phi(x) \subseteq \Phi(x)+\Phi(x) \subseteq \Psi(x) \subseteq \Phi(x)
$$

Therefore, in both cases, the required equalities are true.
Remark 14 Note that if in particular $F$ is as in Example 2, then $0 \in \Phi(A)$ holds for all $A \subseteq X$. Therefore, the above theorem can be applied.

Now, by using a somewhat more complicated argument, we can also prove
Theorem 13 If F is supertriangular, U has a one-sided zero element 0 and $x, y \in X$ are such that

$$
0 \in F(x, y) \cap F(y, x)
$$

then
(1)
$\Phi(x)=\Psi(x)=F(x, y)=S(x, y)$
(2) $\Phi(x)=\Phi(x)+\Phi(y)$.

Proof. If 0 is a right zero element of U , then by using Theorem 11 we can see that

$$
\begin{aligned}
\Phi(x) & =\Phi(x)+\{0\} \subseteq \Phi(x)+F(x, y) \subseteq F(x, y)=F(x, y)+\{0\} \\
& \subseteq F(x, y)+F(y, x)=F(x, y)+R(x, y)=S(x, y) \subseteq \Psi(x) \subseteq \Phi(x) .
\end{aligned}
$$

While, if 0 is a left zero element of U , then we can quite similarly obtain

$$
\begin{aligned}
\Phi(x) & =\{0\}+\Phi(x) \subseteq F(y, x)+\Phi(x) \subseteq F(y, x)=\{0\}+F(y, x) \\
& \subseteq F(x, y)+F(y, x)=F(x, y)+R(x, y)=S(x, y) \subseteq \Psi(x) \subseteq \Phi(x) .
\end{aligned}
$$

Therefore, in both cases, assertion (1) is true.
Now, assertion (2) can be easily derived from assertion (1), by noticing that

$$
\Phi(x)=S(x, y)=F(x, y)+R(x, y)=F(x, y)+F(y, x)=\Phi(x)+\Phi(y) .
$$

From this theorem, it is clear that in particular we also have the following
Corollary 3 If F is supertriangular and U has a one-sided zero element 0 such that $0 \in \mathrm{~F}(\mathrm{x}, \mathrm{y})$ for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, then for any $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ we have
(1) $\Phi(x)=\Psi(x)=F(x, y)=S(x, y)$;
(2) $\Phi(x)=\Phi(x)+\Phi(y)$.

## 10 The particular case when $U$ is a group

By using an argument of Frege [15, 16] and Sincov [36, 23], we can prove
Theorem 14 If F is a nonpartial, triangular function and U is a group, then there exists a function $\xi$ of X to U such that

$$
F(x, y)=\xi(x)-\xi(y)
$$

for all $x, y \in X$.
Proof. By choosing $z \in X$, and defining

$$
\xi(x)=F(x, z)
$$

for all $x \in X$, we can see that

$$
F(x, y)+\xi(y)=F(x, y)+F(y, z)=F(x, z)=\xi(x),
$$

and thus $F(x, y)=\xi(x)-\xi(y)$ for all $x, y \in X$.

Remark 15 If $F$ is nonpartial and supertriangular and $U$ is a group, then by using a similar argument we can only prove that

$$
F(x, y) \subseteq \bigcap_{z \in X}(F(x, z)-F(y, z))
$$

for all $x, y \in X$.
Now, analogously to [38, Theorem 1] of Wilhelmina Smajdor, we can also prove

Theorem 15 If F is nonpartial and supertriangular, U is a commutative group and $\phi$ is a triangular selection function of F , then

$$
F(x, y)=\phi(x, y)+\Phi(x)
$$

for all $x, y \in X$.
Proof. Define

$$
G(x, y)=-\phi(x, y)+F(x, y)
$$

for all $x, y \in X$.
Then, because of $\phi(x, y) \in F(x, y)$, we evidently have

$$
0=-\phi(x, y)+\phi(x, y) \in-\phi(x, y)+F(x, y)=G(x, y)
$$

for all $x, y \in X$. Moreover, by using the assumed triangularity properties of $\phi$ and $F$, we can easily see that

$$
\begin{aligned}
& G(x, y)+G(y, z)=-\phi(x, y)+F(x, y)-\phi(y, z)+F(y, z)= \\
& \quad-(\phi(x, y)+\phi(y, z))+F(x, y)+F(y, z) \subseteq-\phi(x, z)+F(x, z)=G(x, z)
\end{aligned}
$$

for all $x, y, z \in X$.
Hence, by using Corollary 3 and the simple observation that

$$
\phi(x, x)+\phi(x, x)=\phi(x, x)
$$

and thus $\phi(x, x)=0$ for all $x \in X$, we can already infer that

$$
\mathrm{G}(x, y)=\mathrm{G}(x, x)=-\phi(x, x)+F(x, x)=\Phi(x)
$$

and thus

$$
-\phi(x, y)+F(x, y)=\Phi(x)
$$

for all $x, y \in X$. Therefore, the required equality is also true.

Remark 16 It can be easily seen that a converse of Theorem 14 is also true. Therefore, if $F$ is nonpartial and $U$ is a group, then to find a triangular selection function $\phi$ of $F$, it is enough to find only a function $\xi$ of $X$ to $U$ such that

$$
\xi(x)-\xi(y) \in F(x, y)
$$

for all $x, y \in X$.

## 11 The particular case when U is a commutative groupoid

Theorem 16 If F is supertriangular and U is commutative, then R is also supertriangular.

Proof. By Definitions 6 and 7 and the commutativity of $\mathbb{U}$, we have

$$
\begin{aligned}
R(x, y)+R(y, z) & =F(y, x)+F(z, y) \\
& =F(z, y)+F(y, x) \subseteq F(z, x)=R(x, z)
\end{aligned}
$$

for all $x, y, z \in X$.

Theorem 17 If U is commutative, then for any $\mathrm{x}, \mathrm{y}, \mathrm{z} \in \mathrm{X}$ we have
(1) $S(x, y)=S(y, x)$;
(2) $S(x, y) \subseteq \Psi(x) \cap \Psi(y)$.

Proof. By Definition 7 and the commutativity of $U$, we have

$$
\begin{aligned}
S(x, y) & =F(x, y)+R(x, y)=R(y, x)+F(y, x) \\
& =F(y, x)+R(y, x)=S(y, x)
\end{aligned}
$$

Moreover, by the definition of $\Psi$, it is clear that $S(x, y) \subseteq \Psi(x)$. Hence, by using the above symmetry property of $S$, we can already infer that

$$
S(x, y)=S(y, x) \subseteq \Psi(y)
$$

and thus $S(x, y) \subseteq \Psi(x) \cap \Psi(y)$ also holds.

Remark 17 Thus, if U is commutative, then S is already pointwise symmetric in the sense that $S(x, y)=S(y, x)$ for all $x, y \in X$.

Now, concerning the relation $S$, we can also prove the following

Theorem 18 If F is supertriangular and U is a commutative semigroup, then $S$ is also supertriangular.

Proof. By using Definition 7, Theorem 16 and the commutativity and associativity of $U$, we can see that

$$
\begin{aligned}
& S(x, y)+S(y, z)=F(x, y)+R(x, y)+F(y, z)+R(y, z) \\
& =F(x, y)+F(y, z)+R(x, y)+R(y, z) \\
& \subseteq \mathrm{F}(\mathrm{x}, z)+\mathrm{R}(\mathrm{x}, z)=\mathrm{S}(\mathrm{x}, z)
\end{aligned}
$$

for all $x, y, z \in X$.

## 12 The particular case when F is pointwise symmetric

In addition to Theorem 17, we can also prove the following
Theorem 19 If $x, y \in X$ such that $F(x, y)=F(y, x)$, then
(1) $R(x, y)=F(x, y)$;
(2) $S(x, y)=S(y, x)$;
(3) $S(x, y)=F(x, y)+F(x, y)$;
(4) $2 F(x, y) \subseteq S(x, y) \subseteq \Psi(x) \cap \Psi(y)$.

Proof. By Definition 7 and the assumed symmetry property of F, we have

$$
R(x, y)=F(y, x)=F(x, y)
$$

and thus also

$$
S(x, y)=F(x, y)+R(x, y)=F(x, y)+F(x, y)
$$

Thus, assertions (1) and (3) are true.
Now, we can also easily see that

$$
S(y, x)=F(y, x)+F(y, x)=F(x, y)+F(x, y)=S(x, y)
$$

Therefore, assertion (2) is also true.
Hence, as in the proof of Theorem 17, we can already infer that

$$
S(x, y) \subseteq \Psi(x) \cap \Psi(y)
$$

Therefore, to complete the proof of assertion (4), it remains to note only that now

$$
2 F(x, y) \subseteq F(x, y)+F(x, y)=S(x, y)
$$

is also true.

Remark 18 Thus, not only the commutativity of $U$, but the pointwise symmetry of $F$ also implies the pointwise symmetry of $S$.

By [8], in addition to the pointwise symmetry of $F$, one may also naturally investigate the case when $F$ is only weightable in the sense that

$$
w(x)+F(x, y)=R(x, y)+w(y)
$$

for all $x, y \in X$ and some function (or relation) $w$ on $X$ to $U$.
However, it is now more important to note that, as an immediate consequence of our former results, we can also state

Corollary 4 If F is supertriangular and U is commutative, then for any $x, y \in X$ we have

$$
2 S(x, y) \subseteq S(x, y)+S(y, x) \subseteq S(x, x) \cap S(y, y)
$$

Remark 19 Note that the latter corollary only needs the important consequence of the assumed inclusion property of $F$ that $F(x, y)+F(y, x) \subseteq F(x, x)$ for all $x, y \in X$.

In Theorem 11, by using Definition 7, the latter property has been reformulated in the shorter form that $\Psi(x) \subseteq \Phi(x)$ for all $x \in X$. Now, this already implies that $\Psi$ is a selection relation of $\Phi$. Namely, if $x \in X$ such that $\Phi(x) \neq \emptyset$, then because of $\Phi(x)+\Phi(x) \subseteq \Psi(x)$, we also have $\Psi(x) \neq \emptyset$.

## 13 The particular case when $U$ is a group and $F$ is pointwise skew symmetric

Analogously to Theorem 19, we can also prove the following
Theorem 20 If U is a group and $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ such that $\mathrm{F}(\mathrm{x}, \mathrm{y})=-\mathrm{F}(\mathrm{y}, \mathrm{x})$, then
(1) $R(x, y)=-F(x, y)$;
(2) $S(x, y)=-S(y, x)$;
(3) $S(x, y)=F(x, y)-F(x, y)$;
(4) $S(x, y) \subseteq \Psi(x) \cap(-\Psi(y))$.

Proof. To prove (4), note that now, in addition to $S(x, y) \subseteq \Psi(x)$, we also have

$$
S(x, y)=-S(y, x) \subseteq-\Psi(y),
$$

and thus $S(x, y) \subseteq \Psi(x) \cap(-\Psi(y))$ also holds.
Remark 20 If in addition to the assumptions of this theorem $F(x, y) \neq \emptyset$ also holds, then from assertion (3) we can infer that $0 \in S(x, y)$.

Now, by using the corresponding definitions and Theorem 20, we can also prove

Theorem 21 If U is a group and F is pointwise skew symmetric, then for any $x \in X$ we have
(1) $\Phi(x)=-\Phi(x)$;
(2) $\Psi(x)=-\Psi(x)$.

Proof. To prove (2), note that by Definition 7 and Theorem 20 we have

$$
\Psi(x)=\bigcup_{y \in X} S(x, y)=\bigcup_{y \in X}(-S(x, y))=-\bigcup_{y \in X} S(x, y)=-\Psi(x)
$$

for all $x \in X$.
Remark 21 If in addition to the assumptions of this theorem, $\Phi(x) \neq \emptyset$ also holds, then from the inclusion

$$
\Phi(x)-\Phi(x)=\Phi(x)+\Phi(x) \subseteq \Psi(x)
$$

we can infer that $0 \in \Psi(x)$. Therefore, if in addition $F$ is supertriangular, then because Theorem 11, we also have $0 \in \Phi(x)$.

Thus, by Theorem 12, we can also state the following
Theorem 22 If U is a group and F is nonpartial, supertriangular and pointwise skew symmetric, then for any $\mathrm{x} \in \mathrm{X}$ we have
(1) $\Phi(x)=\Psi(x)$;
(2) $\Phi(x)=\Phi(x)+\Phi(x)$.

Now, by Theorems 20 and 21, we can also state the following
Theorem 23 If U is a group and F is a nonpartial, pointwise skew symmetric function, then for any $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ we have
(1) $S(x, y)=0$;
(2) $\Phi(x)=\Psi(x)=0$.

The following example shows the three important consequences of the inclusion considered in Definition 6 do not imply, even in a very simple case, the validity of this inclusion itself.

## Example 3 If

$$
F(x, y)=\operatorname{sgn}(x-y)
$$

for all $\mathrm{x}, \mathrm{y} \in \mathbb{R}$, then F is a skew symmetric function of $\mathbb{R}^{2}$ to $\mathbb{R}$ such that, under the notation $\Phi(x)=F(x, x)$, for any $x, y \in X$ we have
(1) $F(x, y)+F(y, x)=\Phi(x)$;
(2) $\Phi(x)+F(x, y)=F(x, y)$;
(3) $F(x, y)+\Phi(y)=F(x, y)$.

However, F is not either supertriangular nor subtriangular in both functional and relational sense.

Namely, for instance, we have

$$
\mathrm{F}(2,1)+\mathrm{F}(1,0)=2 \quad \text { and } \quad \mathrm{F}(2,0)=1
$$

and

$$
\mathrm{F}(0,1)+\mathrm{F}(1,2)=-2 \quad \text { and } \quad \mathrm{F}(0,2)=-1 .
$$

## 14 The particular case when U is cancellative

Definition 8 In what follows, we shall denote by lcan(U) and rcan(U) the family of all left-cancellable and right-cancellable elements of the groupoid U , respectively.

Moreover, we shall also write $\operatorname{can}(\mathrm{U})=\operatorname{lcan}(\mathrm{U}) \cap \operatorname{rcan}(\mathrm{U})$.
Remark 22 Thus, for any $u \in \mathrm{U}$, we have $\boldsymbol{u} \in \operatorname{lcan}(\mathrm{U})$ if and only if $u+v=$ $u+w$ implies $\mathfrak{u}=w$ for all $v, w \in \mathbf{U}$.

Moreover, for instance, we can state that U is left-cancellative if and only if $\operatorname{lcan}(\mathrm{U})=\mathrm{U}$.

Lemma 1 For any $\mathrm{V}, \mathrm{W} \subseteq \mathrm{U}$,
(1) $\operatorname{card}(\mathrm{V}+\mathrm{W}) \leq 1$ and $\mathrm{V} \cap \operatorname{lcan}(\mathrm{U}) \neq \emptyset$ imply that $\operatorname{card}(\mathrm{W}) \leq 1$;
(2) $\operatorname{card}(\mathrm{V}+\mathrm{W}) \leq 1$ and $\mathrm{W} \cap \operatorname{rcan}(\mathrm{U}) \neq \emptyset$ imply that $\operatorname{card}(\mathrm{V}) \leq 1$.

Proof. Assume that the conditions of (1) hold, $v \in \mathrm{~V} \cap \operatorname{lcan}(\mathrm{U})$ and $w_{1}, w_{2} \in$ $W$. Then, we have $v+w_{1}, v+w_{2} \in \mathrm{~V}+\mathrm{W}$. Hence, by using that $\operatorname{card}(\mathrm{V}+\mathrm{W}) \leq$ 1 , we can infer that $v+w_{1}=v+w_{2}$. Moreover, since $v \in \operatorname{lcan}(\mathrm{U})$, we can also state that $w_{1}=w_{2}$. Therefore, $\operatorname{card}(W) \leq 1$, and thus (1) also holds.

The proof of assertion (2) is quite similar.
Now, by using this lemma, we can give some reasonable sufficient conditions in order that a suppertriangular relation should be a function.

Theorem 24 If F is supertriangular and there exist $\mathrm{x}_{0}, \mathrm{y}_{0} \in \mathrm{X}$ such that
(1) $\operatorname{card}\left(F\left(x_{0}, y_{0}\right)\right) \leq 1$;
(2) $\mathrm{F}\left(\mathrm{x}, \mathrm{y}_{0}\right) \cap \operatorname{rcan}(\mathrm{U}) \neq \emptyset$ for all $\mathrm{x} \in \mathrm{X}$;
(3) $\mathrm{F}\left(\mathrm{x}_{0}, \mathrm{y}\right) \cap \operatorname{lcan}(\mathrm{U}) \neq \emptyset$ for all $\mathrm{y} \in \mathrm{X}$;
then $\operatorname{card}(\mathrm{F}(\mathrm{x}, \mathrm{y})) \leq 1$ for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, and thus F is a function.
Proof. By the assumed inclusion property of $F$, we have

$$
F\left(x_{0}, x\right)+F\left(x, y_{0}\right) \subseteq F\left(x_{0}, y_{0}\right)
$$

for all $x \in X$. Hence, by using conditions (1) and (3) and Lemma 1, we can infer that
(a) $\operatorname{card}\left(F\left(x, y_{0}\right)\right) \leq 1$ for all $x \in X$.

Now, by the assumed inclusion property of $F$, we also have

$$
F(x, y)+F\left(y, y_{0}\right) \subseteq F\left(x, y_{0}\right)
$$

for all $x, y \in X$. Hence, by using assertion (a) condition (2) and Lemma 1, we can infer that
(b) $\operatorname{card}(F(x, y)) \leq 1$ for all $x, y \in X$.

Thus, the required assertion is true.
From this theorem, by using Theorem 14, we can immediately derive
Corollary 5 If F is nonpartial and supertriangular, U is a group and $\operatorname{card}\left(\mathrm{F}\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right)\right)=1$ for some $\mathrm{x}_{0}, \mathrm{y}_{0} \in \mathrm{X}$, then there exists a function $\xi$ of X to U such that

$$
F(x, y)=\xi(x)-\xi(y)
$$

for all $x, y \in X$.

## 15 The particular case when $U$ has a suitable distance function

Remark 23 A function $d$ of $X^{2}$ to $[0,+\infty]$ is usually called a distance function on $X$.

Moreover, the extended real number

$$
d(X)=\operatorname{diam}(X)=\sup \{d(x, y): \quad x, y \in X\}
$$

is called the diameter of X .
Remark 24 Thus, we have $d(X)=-\infty$ if $X=\emptyset$, and $d(X) \geq 0$ if $X \neq \emptyset$. Moreover, if $X \neq \emptyset$, then $\operatorname{card}(X)=+\infty$ may also hold even if $X$ is finite.

Definition 9 A distance function d on X will be called admissible if
(a) $d(X)<+\infty$;
(b) $\mathrm{d}(\mathrm{x}, \mathrm{y})=0$ implies $\mathrm{x}=\mathrm{y}$ for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$.

Moreover, the distance function d will be called extremal if
(c) for any $x, y \in X$ there exist $c \in] 1,+\infty[$ and $z, w \in X$ such that

$$
\operatorname{cd}(x, y) \leq d(z, w)
$$

Remark 25 If X is an additive groupoid, then to satisfy condition (c) we may naturally assume that for any $x, y \in X$, there exists $n \in \mathbb{N} \backslash\{1\}$ such that

$$
n d(x, y) \leq d(n x, n y)
$$

Namely, if X is a commutative abelian group and p is a function of U to $[0,+\infty]$ such that

$$
n p(x) \leq p(n x)
$$

for all $n \in \mathbb{N}$ and $x \in X$, then by defining

$$
d(x, y)=p(-x+y)
$$

for all $x, y \in X$, we have

$$
\begin{aligned}
\operatorname{nd}(x, y) & =n p(-x+y) \leq p(n(-x+y)) \\
& =p(n(-x)+n y)=p(-n x+n y)=d(n x, n y)
\end{aligned}
$$

for all $\mathrm{n} \in \mathbb{N}$ and $x, \mathrm{y} \in \mathrm{U}$.
The introduction of Definition 9 can only be motivated by the following
Lemma 2 If there exists an extremal, admissible distance function d on X , then $\operatorname{card}(\mathrm{X}) \leq 1$.

Proof. If $X=\emptyset$, then the required assertion trivially holds. Therefore, we may assume that $X \neq \emptyset$, and thus $d(X) \neq-\infty$. Now, by condition (a), we can state that $d(X) \in \mathbb{R}$. Moreover, since $d$ is nonnegative, we can now also note that $\mathrm{d}(\mathrm{X}) \geq 0$.

Thus, for every $\varepsilon>0$, we have

$$
\mathrm{d}(\mathrm{X})-\varepsilon<\mathrm{d}(\mathrm{X}) .
$$

Therefore, by the definition of $d(X)$, there exist $x, y \in X$ such that $d(X)-\varepsilon<d(x, y)$, and thus

$$
\mathrm{d}(\mathrm{X})<\mathrm{d}(\mathrm{x}, \mathrm{y})+\varepsilon .
$$

Moreover, by condition (c), there exist $c \in] 1,+\infty[$ and and $z, w \in X$ such that

$$
\operatorname{cd}(x y) \leq d(z, w)
$$

Combining the above two inequalities, we can see that

$$
\operatorname{cd}(x, y)<d(z, w) \leq d(X)<d(x, y)+\varepsilon
$$

and thus $(c-1) d(x, y)<\varepsilon$. Hence, by letting $\varepsilon$ tend to zero, we can infer that $(c-1) d(x, y) \leq 0$. Therefore, since $c-1>0$, we necessarily have $d(x, y) \leq 0$, and hence $\mathrm{d}(\mathrm{x}, \mathrm{y}) \leq 0$ by the nonnegativity of d . Thus, we actually have

$$
\mathrm{d}(X)<\mathrm{d}(x, x)+\varepsilon=\varepsilon .
$$

Hence, by letting $\varepsilon$ tend to zero, we can infer that $d(X) \leq 0$, and thus also $d(X)=0$ by the nonnegativity of $d(X)$.
This, by condition (b), already implies that $\operatorname{card}(X)=1$. Namely, if this is not the case, then by the assumption $X \neq \emptyset$, there exist $x, y \in X$ such that $x \neq y$. Hence, by condition (b) and the nonnegativity of $d$, we can infer that $d(x, y)>0$, and thus also $d(X)>0$ by the definition of $d(X)$. This contradiction proves that $\operatorname{card}(X)=1$.

Remark 26 From condition (c), by induction, we can infer that there exist sequences $\left(c_{n}\right)_{n=1}^{\infty}$ in $] 1,+\infty\left[\right.$ and $\left(x_{n}\right)_{n=1}^{\infty}$ and $\left(y_{n}\right)_{n=1}^{\infty}$ in $X$ such that

$$
d(x, y) \prod_{i=0}^{n} c_{i} \leq d\left(x_{n}, y_{n}\right)
$$

for all $n \in \mathbb{N}$. However, this fact cannot certainly be used to give a simpler proof for Lemma 2.

From Theorem 24, by using Lemma 2, we can immediately derive
Theorem 25 If F is supertriangular and there exist $\mathrm{x}_{0}, \mathrm{y}_{0} \in \mathrm{X}$, such that
(1) $\mathrm{F}\left(\mathrm{x}, \mathrm{y}_{0}\right) \cap \operatorname{rcan}(\mathrm{U}) \neq \emptyset$; for all $\mathrm{x} \in \mathrm{X}$;
(2) $\mathrm{F}\left(\mathrm{x}_{0}, \mathrm{y}\right) \cap \operatorname{lcan}(\mathrm{U}) \neq \emptyset$; for all $\mathrm{y} \in \mathrm{X}$;
(3) there exists an extremal, admissible distance function on $\mathrm{F}\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right)$; then $\operatorname{card}(\mathrm{F}(\mathrm{x}, \mathrm{y})) \leq 1$ for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, and thus F is a function.

Proof. By assumption (3) and Lemma 2, we have $\operatorname{card}\left(F\left(x_{0}, y_{0}\right)\right) \leq 1$. Hence, by Theorem 24, we can see that the required assertion is also true.

## 16 Contructions of supertriangular relations

Theorem 26 If V is a subgroupoid of U and

$$
F(x, y)=V
$$

for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, then F is a supertriangular relation on X to U .
Proof. We evidently have

$$
\mathrm{F}(\mathrm{x}, \mathrm{y})+\mathrm{F}(\mathrm{y}, \mathrm{z})=\mathrm{V}+\mathrm{V} \subseteq \mathrm{~V}=\mathrm{F}(\mathrm{x}, \mathrm{z})
$$

for all $x, y, z \in X$.
Remark 27 Conversely, note that if $F$ is a supertriangular relation on $X^{2}$ to U , then by Corollary $2 \Phi(x)=F(x, x)$ is a subgroupoid of $U$ for all $x \in X$.

Now, as a converse to Theorem 14, we can also easily prove the following
Theorem 27 If $\xi$ is a function of X to $\mathrm{U}, \mathrm{U}$ is a group and

$$
F(x, y)=\xi(x)-\xi(y)
$$

for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, then F is a triangular function of $\mathrm{X}^{2}$ to U .
Proof. We evidently have

$$
F(x, y)+F(y, z)=\xi(x)-\xi(y)+\xi(y)-\xi(z)=\xi(x)-\xi(z)=F(x, z)
$$

for all $x, y, z \in X$.

Remark 28 If $\xi$ is only a relation of $X$ to $U, U$ is a group and $F(x, y)=$ $\xi(x)-\xi(y)$ for all $x, y \in X$, then by using a similar argument we can only prove that $F$ is a subtriangular relation of $X^{2}$ to $U$.

In addition to the above two theorems, it is also worth proving that the family of all supertriangular relations is closed under the usual pointwise operations.

Theorem 28 If F is a supertriangular relation on $\mathrm{X}^{2}$ to U and U is a commutative semigroup, then nF is also a supertriangular relation on $\mathrm{X}^{2}$ to U for all $\mathrm{n} \in \mathbb{N}$.

Proof. If $n \in \mathbb{N}$, then by the corresponding definitions we have

$$
\begin{aligned}
(n F)(x, y)+(n F)(y, z) & =n F(x, y)+n F(y, z) \\
& =n(F(x, y)+F(y, z)) \subseteq n F(x, z)=(n F)(x, z)
\end{aligned}
$$

for all $x, y, z \in X$.
Remark 29 If $F$ is a supertriangular relation on $X^{2}$ to $U$ and $U$ has a zero element, then

$$
(O F)(x, y)=\emptyset \quad \text { if } \quad F(x, y)=\emptyset \quad \text { and } \quad(O F)(x, y)=\{0\} \quad \text { if } \quad F(x, y) \neq \emptyset
$$

Therefore, $O F$ is a supertriangular function on $X^{2}$ to $U$.
Now, analogously to Theorem 28, we can also prove the following
Theorem 29 If F is a supertriangular relation on $\mathrm{X}^{2}$ to U and U is a commutative group, then kF is also a supertriangular relation on $\mathrm{X}^{2}$ to U for all $k \in \mathbb{Z}$.

Moreover, in addition to Theorems 28, we can also easily prove the following
Theorem 30 If F and G are supertriangular relations on $\mathrm{X}^{2}$ to U and U is a commutative semigroup, then $\mathrm{F}+\mathrm{G}$ is also a supertriangular relation on $\mathrm{X}^{2}$ to U.

Proof. By the corresponding definitions, it is clear that

$$
\begin{aligned}
& (F+G)(x, y)+(F+G)(y, z)=F(x, y)+G(x, y)+F(y, z)+G(y, z) \\
& =F(x, y)+F(y, z)+G(x, y)+G(y, z) \subseteq F(x, z)+G(x, z)=(F+G)(x, z)
\end{aligned}
$$

for all $x, y, z \in X$.

## 17 An application of the above results

Now, by using Theorems 26, 27 and 30, we can also easily establish
Theorem 31 If $\xi$ a function of X to $\mathrm{U}, \mathrm{U}$ is a commutative group, V is a subgroupoid of U and

$$
F(x, y)=\xi(x)-\xi(y)+V
$$

for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, then F is a supertriangular relation on $\mathrm{X}^{2}$ to U such that, under the notations of Definition 7, for any $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ we have:
(1) $\Phi(x)=V$;
(2) $\mathrm{S}(\mathrm{x}, \mathrm{y})=\mathrm{V}+\mathrm{V}$;
(3) $\Psi(x)=\emptyset$ if $\mathrm{X}=\emptyset$ and $\Psi(\mathrm{x})=\mathrm{V}+\mathrm{V}$ if $\mathrm{X} \neq \emptyset$.

Proof. From Theorems 26, 27 and 30, it is clear that F is supertriangular. Moreover, by the corresponding definitions, it is clear that

$$
\begin{gathered}
\Phi(x)=F(x, x)=\xi(x)-\xi(x)+V=V \\
S(x, y)=F(x, y)+F(y, x)=\xi(x)-\xi(y)+V+\xi(y)-\xi(x)+V=V+V
\end{gathered}
$$

and

$$
\Psi(x)=\bigcup_{y \in X} S(x, y)=\bigcup_{y \in X}(V+V)=\left\{\begin{array}{lll}
\emptyset & \text { if } & X=\emptyset \\
V+V & \text { if } & X \neq \emptyset
\end{array}\right.
$$

Moreover, for an easy illustration of this theorem, we can also state
Example 4 If $\mathrm{r} \geq 0$ and

$$
F(x, y)=[x-y+r,+\infty[
$$

for all $\mathrm{x}, \mathrm{y} \in \mathbb{R}$, then F is a supertriangular relation of $\mathbb{R}^{2}$ to $\mathbb{R}$ such that, for any $\mathrm{x}, \mathrm{y} \in \mathrm{X}$, we have:
(1) $\Phi(\mathrm{x})=[\mathrm{r},+\infty[$;
(2) $\Psi(x)=S(x, y)=[2 r,+\infty[$.

To check this, note that, by taking $\xi=\Delta_{\mathbb{R}}$ and $\mathrm{V}=[\mathrm{r},+\infty[$, we have

$$
F(x, y)=[x-y+r,+\infty[=x-y+[r,+\infty[=\xi(x)-\xi(y)+V
$$

for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$. Therefore, Theorem 31 can be applied.
For instance, by assertion (2) of Theorem 31, we have

$$
S(x, y)=V+V=[r,+\infty[+[r,+\infty[=[2 r,+\infty[
$$

for all $x, y \in X$.

Remark 30 Note that in the present particular case, for any $x, y \in \mathbb{R}$, we have:
(1) $0 \in \Phi(x) \Longleftrightarrow r=0$;
(2) $\Phi(x)=\Psi(x) \Longleftrightarrow \mathrm{r}=0$;
(3) $x-y \in F(x, y) \Longleftrightarrow r=0$;
(4) $0 \in F(x, y) \Longleftrightarrow r \leq y-x$;
(5) $0 \in F(x, y) \cap F(y, x) \Longleftrightarrow r=0, \quad x=y$.

To prove (5), note that by (4) we have
$0 \in F(x, y) \cap F(y, x) \Longleftrightarrow r \leq y-x, r \leq x-y \Longleftrightarrow r \leq \min \{x-y, y-x\}$.
Moreover, recall that $\min \{a, b\}=2^{-1}(a+b-|a-b|)$ for all $a, b \in \mathbb{R}$, and thus in particular $\min \{x-y, y-x\}=-|x-y|$. Therefore,

$$
r \leq \min \{x-y, y-x\} \Longleftrightarrow r \leq-|x-y| \Longleftrightarrow r=0, \quad x=y
$$

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## References

[1] J. Aczél, Lectures on Functional Equations and Their Applications, Academic Press, New York, 1966.
[2] M. Alimohammady, S. Jafari, S. P. Moshokoa and M. K. Kalleji, A note on properties of hypermetric spaces, J. Hyperstructures, 3 (2014), 89-100.
[3] P. Augustová and L. Klapka, Atlas as solutions of Sincov's inequality, arXiv: 1612.00355v1 [math.DS] 1 Dec 2016, 7 pp.
[4] J. A. Baker, Solution of problem E 2607, Amer. Math. Monthly, 84 (1977), 824-825.
[5] G. Birkhoff, Lattice Theory, Amer. Math. Soc. Colloq. Publ., 25, Providence, Rhode Island, 1967.
[6] H. Brandt, Über eine Verallgemeinerung der Gruppenbregriffes, Math. Ann., 96 (1926), 360-366.
[7] M. J. Campión, R. G. Catalán, E. Induráin and G. Ochoa, Reinterpreting a fuzzy subset by means of a Sincov's functional equation, J. Intelligent Fuzzy Systems, 27 (2014), 367-375.
[8] M. J. Campión, E. Induráin, G. Ochoa and O. Valero, Functional equations related to weightable quasi-metrics, Hacet. J. Math. Stat., 44 (2015), 775-787.
[9] M. Cantor, Funktionalgleichungen mit drei von einander unabhängigen Veränderlichen, Zeitschrift Mat. Physik., 41 (1896), 161-163.
[10] E. Castillo-Ron and R. Ruiz-Cobo, Functional Equations in Science and Engineering, Marcel Decker, New York, 1992.
[11] R. Croisot, Une interprétation des relations d'équivalence dans un ensemble, C. R. Acad. Sci. Paris, 226 (1948), 616-617.
[12] B. A. Davey and H. A. Priestley, Introduction to Lattices and Order, Cambridge University Press, Cambridge, 2002.
[13] W. Fechner, Richard's inequality, Cauchy-Schwarz's inequality and approximate solutions of Sincov's equation, Proc. Amer. Math. Soc., 147 (2019), 3955-3960.
[14] W. Fechner, Sincov's inequalities on topological spaces, Publ. Math. Debrecen, 96 (2020), 63-76.
[15] G. Frege, Rechnungsmethoden, die sich auf eine Erweiterung des Grössenbegriffes gründen, Dissertation for the Venia docendi, Verlag Friedrich Frommann, Jena, 1874.
[16] G. Frege, Methods of calculation based on an extension of the concept of quantity, In: G. Frege, Collected Papers of on Mathematics, Logic, and Philosophy, Basil Blackwell, Oxford, 1984, 56-92.
[17] Z. Gajda, Invariant means and representations of semigroups in the theory of functional equations, Prace Naukowe Uniwersytetu Śla̧skiego w Katowicach, 1273 (1992), 1-81.
[18] T. Glavosits, Generated preorders and equivalences, Acta Acad. Paed. Agriensis, Sect. Math., 29 (2002), 95-103.
[19] T. Glavosits and Á. Száz, Constructions and extensions of free and controlled additive relations, In: Th.M. Rassias (Ed.), Handbook of Functional Equations: Functional Inequalities, Springer Optim. Appl. 95 (2014), 161-208.
[20] D. Gronau, Gottlob Frege, a pioneer in iteration theory, In: L. Reich, J. Smítal and Gy. Targonski (Eds.), Proceedings of the European Conference on Iteration Theory, ECIT94, Gracer Math. Ber. 334 (1997), 105-119.
[21] D. Gronau, Gottlob Fregees Beiträge zur Iteratiostheorie und zur Theorie der Functionalgleichungen, In: G. Gabriel (Ed.), Gottlob Frege - Werk und Wirkung, Mentis Verlag, Paderborn, 200, 151-169.
[22] D. Gronau, A remark on Sincov's functional equation, Notices South African Math. Soc. Esaim: Proceeding and Surveys, 31 (2000), 1-8.
[23] D. Gronau, Translation equation and Sincov's equation - A historical remark, Esaim: Proceeding and Surveys 46 (2014), 43-46.
[24] J. Mala and Á. Száz, Modifications of relators, Acta Math. Hungar., 77 (1997), 69-81.
[25] Z. Moszner, Solution générale de l'équation $F(x, y) F(y, z)=F(x, z)$ pour $x \leq y \leq z, C . R$. Acad. Sci. Paris, 261 (1965), 28.
[26] Z. Moszner, L'équation de translation et l1équation de Sincov généralisée, Rocznik Nauk.-Dydakt. Prace Mat., 16 (1999), 53-71.
[27] Z. Moszner, On the stability of functional equations, Aequationes Math., 77 (2009), 33-88.
[28] K. Nikodem, Additive selections of additive set-valued functions, Zb. Rad. Prirod.-Mat. Fak., 18 (1988), 143-148.
[29] K. Nikodem and D. Popa, On single-valuedness of set-valued maps satisfying linear inclusions, Banach J. Math. Anal., 3 (2009), 44-51.
[30] G. Pataki and Á. Száz, A unified treatment of well-chainedness and connectedness properties, Acta Math. Acad. Paedagog. Nyházi. (N.S.), 19 (2003), 101-165.
[31] J. Pepis, Sur une famille d'ensembles plans et les solutions de l'équation fonctionnelle $F(x, z)=F(x, y) \cdot F(y, z)$ pour $0 \leq x \leq y \leq z$. Application á la théorie générale des intéréts, Ann. Soc. Polon. Math., 17 (1937), 113.
[32] W. J. Pervin, Quasi-uniformization of topological spaces, Math. Ann., 147 (1962), 316-317.
[33] B. Piatek, On the Sincov functional equation, Demonstratio Math., 38 (2005), 875-881.
[34] P. K. Sahoo, Stability of a Sincov type functional equation, J. Inf. Math. Sci., 1 (2009), 81-90.
[35] P. K. Sahoo, On a Sincov type functional equation, In: Th.M. Rassias and J. Brzdek (Eds.), Functional Equations in Mathematical Analysis, Springer Optimizations and Its Applications, 52, Chapter 43, Springer, New York, 2012, 697-7008.
[36] D. M. Sincov, Über eine Funktionalgleichung, Arch. Math. Phys., 6 (1904), 216-217.
[37] A. Smajdor, Iterations of multi-valued functions, Prace Naukowe Universytetu Ślaskiego w Katowicach 759, Universitet Ślaski, Katowice 1985.
[38] W. Smajdor, Set-valued version of Sincov's functional equation, Demonstration Math., 39 (2006), 101-105.
[39] Á. Száz, Galois type connections and closure operations on preordered sets, Acta Math. Univ.Comen., 78 (2009), 1-21.
[40] Á. Száz, Corelations are more powerful tools than relations, In: Th. M. Rassias (Ed.), Applications of Nonlinear Analysis, Springer Optimization and Its Applications, 134 (2018), 711-779.
[41] A. Weil, Sur les espaces á structure uniforme at sur la topologie générale, Actual. Sci. Ind., 551 Herman and Cie, Paris, 1937.

# Direct and converse theorems for King operators 

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#### Abstract

For the sequence of King operators, we establish a direct approximation theorem via the first order Ditzian-Totik modulus of smoothness, and a converse approximation theorem of Berens-Lorentz-type.


## 1 Introduction

Studying the connection between regular summability matrices and convergent positive linear operators, King [3] introduced an interesting Bernstein-type operator defined as follows:

$$
\begin{equation*}
\left(V_{n} f\right)(x) \equiv V_{n}(f ; x)=\sum_{k=0}^{n} p_{n, k}\left(r_{n}(x)\right) f\left(\frac{k}{n}\right), \tag{1}
\end{equation*}
$$

where $x \in[0,1], f \in C[0,1], p_{n, k}(x)=\binom{n}{k} x^{k}(1-x)^{n-k}$ and

$$
r_{n}(x)=\left\{\begin{align*}
x^{2}, & \text { if } n=1  \tag{2}\\
-\frac{1}{2(n-1)}+\sqrt{\frac{n}{n-1} x^{2}+\frac{1}{4(n-1)^{2}},} & \text { if } n=2,3, \ldots
\end{align*}\right.
$$

[^2]For $r_{n}(x)=x, x \in[0,1]$, we recover from (1) the classical Bernstein operator:

$$
\left(B_{n} f\right)(x) \equiv B_{n}(f ; x)=\sum_{k=0}^{n} p_{n, k}(x) f\left(\frac{k}{n}\right)
$$

It is known that

$$
\begin{equation*}
\left(B_{n} e_{0}\right)(x)=1, \quad\left(B_{n} e_{1}\right)(x)=x \quad \text { and } \quad\left(B_{n} e_{2}\right)(x)=x^{2}+\frac{x(1-x)}{n} \tag{3}
\end{equation*}
$$

where $e_{j}(x)=x^{j}, x \in[0,1]$ and $j \in\{0,1,2, \ldots\}$. In contrast with (3), we have for $V_{n}$ the relations (see [3, pp. 204-205]):

$$
\begin{equation*}
\left(V_{n} e_{0}\right)(x)=1, \quad\left(V_{n} e_{1}\right)(x)=r_{n}(x) \quad \text { and } \quad\left(V_{n} e_{2}\right)(x)=x^{2} . \tag{4}
\end{equation*}
$$

The goal of the paper is to obtain direct and converse approximation theorems for the operators given by (1)-(2). The direct result is established with the aid of the first order Ditzian-Totik modulus of smoothness defined by

$$
\begin{equation*}
\omega_{\varphi}^{1}(f ; \delta)=\sup _{0<h \leq \delta} \sup _{x \pm \frac{1}{2} h \varphi(x) \in[0,1]}\left|f\left(x+\frac{1}{2} h \varphi(x)\right)-f\left(x-\frac{1}{2} h \varphi(x)\right)\right| \tag{5}
\end{equation*}
$$

where $\varphi(x)=\sqrt{x(1-x)}, x \in[0,1]$. It is known [2, Theorem 2.1.1] that (5) is equivalent with the K -functional

$$
\mathrm{K}_{1, \varphi}(\mathrm{f} ; \delta)=\inf _{\mathrm{g} \in \mathcal{W}(\varphi)}\left\{\|\mathrm{f}-\mathrm{g}\|+\delta\left\|\varphi \mathrm{g}^{\prime}\right\|\right\}, \quad \delta>0
$$

where $\mathcal{W}(\varphi)=\left\{g \mid g \in A . C .{ }_{\text {loc }}[0,1],\left\|\varphi g^{\prime}\right\|<\infty\right\}$, i.e. there exists $C_{1}>0$ such that

$$
\begin{equation*}
C_{1}^{-1} \omega_{\varphi}^{1}(f ; \delta) \leq K_{1, \varphi}(f ; \delta) \leq C_{1} \omega_{\varphi}^{1}(f ; \delta) \tag{6}
\end{equation*}
$$

Finally, a converse result of Berens-Lorentz-type is established for the operators $\mathrm{V}_{\mathrm{n}}$ (see [1, p. 312, Lemma 5.2] and Lemma 3 below). Throughout this paper $C_{1}, C_{2}, \ldots, C_{13}$ denote absolute positive constants.

## 2 Direct theorem

We have the following result for the functions defined by (2).
Lemma 1 The functions $\mathrm{r}_{\mathrm{n}}, \mathrm{n}=1,2, \ldots$, satisfy the properties
a) $0 \leq r_{n}^{\prime}(x) \leq 2$ for $x \in[0,1]$ and $n=1,2, \ldots$;
b) $r_{n}(0)=0, r_{n}(1)=1$ and $r_{n}$ is strictly increasing function on $[0,1]$ for $n=1,2, \ldots$;
c) $0 \leq r_{n}(x) \leq x \leq 1$ for $x \in[0,1]$ and $n=1,2, \ldots$;
d) $0 \leq x-r_{n}(x) \leq \frac{2}{n}(1-x)$ for $x \in[0,1]$ and $n=1,2, \ldots$;
e) $x \leq 2 r_{n}(x)$ for $x \in\left[\frac{1}{n}, 1\right]$ and $n=1,2, \ldots$

Proof. a) Obviously $r_{1}^{\prime}(x)=2 x, x \in[0,1]$. For $n \geq 2$, by simple computations, we obtain

$$
r_{n}^{\prime}(x)= \begin{cases}\lim _{x \searrow 0} \frac{r_{n}(x)-r_{n}(0)}{x-0}=0, & \text { if } x=0 \\ \frac{\frac{n}{n-1} x}{\sqrt{\frac{n}{n-1} x^{2}+\frac{1}{4(n-1)^{2}}},} & \text { if } 0<x \leq 1\end{cases}
$$

Hence $0 \leq r_{n}^{\prime}(x) \leq \frac{\frac{n}{n-1} x}{\sqrt{\frac{n}{n-1}} x}=\sqrt{\frac{n}{n-1}} \leq \sqrt{2}$ for $x \in(0,1]$. Thus $0 \leq r_{n}^{\prime}(x) \leq 2$ for $n=1,2, \ldots$ and $x \in[0,1]$.
b) It follows from (2) and a).
c) It follows from (2) by direct computations.
d) Obviously $0 \leq x-r_{1}(x)=x(1-x) \leq 2(1-x), x \in[0,1]$. Using $\left.b\right)$ and c), we have $0 \leq x-r_{n}(x) \leq \frac{2}{n}(1-x)$ for $x=0$ and $n \geq 2$, and

$$
\begin{aligned}
0 \leq x-r_{n}(x) & =x+\frac{1}{2(n-1)}-\sqrt{\frac{n}{n-1} x^{2}+\frac{1}{4(n-1)^{2}}} \\
& =\frac{\frac{x(1-x)}{n-1}}{x+\frac{1}{2(n-1)}+\sqrt{\frac{n}{n-1} x^{2}+\frac{1}{4(n-1)^{2}}}} \\
& \leq \frac{\frac{x(1-x)}{n-1}}{x}=\frac{1-x}{n-1} \leq \frac{2}{n}(1-x)
\end{aligned}
$$

for $x \in(0,1]$ and $n \geq 2$.
e) For $n=1$ the statement is obvious. For $n \geq 2$, we consider the function $h(x)=\frac{x}{r_{n}(x)}, x \in\left[\frac{1}{n}, 1\right]$. Then, by (2),

$$
h^{\prime}(x)=\frac{r_{n}(x)-x r_{n}^{\prime}(x)}{r_{n}^{2}(x)}=r_{n}^{-2}(x)\left(r_{n}(x)+\frac{1}{2(n-1)}\right)^{-1 / 2}
$$

$$
\times \frac{1}{2(n-1)}\left[\frac{1}{2(n-1)}-\sqrt{\frac{n}{n-1} x^{2}+\frac{1}{4(n-1)^{2}}}\right]<0
$$

for $x \in\left[\frac{1}{n}, 1\right]$. Hence

$$
\begin{aligned}
h(x) & \leq h\left(\frac{1}{n}\right)=\frac{\frac{1}{n}}{-\frac{1}{2(n-1)}+\sqrt{\frac{n}{n-1} \frac{1}{n^{2}}+\frac{1}{4(n-1)^{2}}}} \\
& =\frac{2(n-1)}{\sqrt{n}} \frac{1}{\sqrt{5 n-4}-\sqrt{n}}=\frac{1}{2} \frac{\sqrt{5 n-4}+\sqrt{n}}{\sqrt{n}} \leq 2
\end{aligned}
$$

for $x \in\left[\frac{1}{n}, 1\right]$, which was to be proved.
The operators $\mathrm{V}_{\mathrm{n}}$ given by (1)-(2) are linear and positive. By Lemma 1, b), we have

$$
\begin{equation*}
\left(V_{n} f\right)(0)=f(0) \quad \text { and } \quad\left(V_{n} f\right)(1)=f(1) \tag{7}
\end{equation*}
$$

for all $f \in C[0,1]$.
In the next theorem we establish the direct result.
Theorem 1 There exists $C_{2}>0$ such that

$$
\begin{equation*}
\left\|V_{n} f-f\right\| \leq C_{2} \omega_{\varphi}^{1}\left(f ; \frac{1}{\sqrt{n}}\right) \tag{8}
\end{equation*}
$$

for all $\mathrm{f} \in \mathrm{C}[0,1]$ and $\mathrm{n}=1,2, \ldots$
Proof. Let $x \in(0,1)$ and $t \in[0,1]$. Taking into account [2, Lemma 9.6.1], we have

$$
\begin{equation*}
\left|\int_{x}^{t} \frac{d u}{\varphi(u)}\right| \leq \varphi^{-1}(x)|t-x|^{1 / 2}\left|\int_{x}^{t} \frac{d u}{|t-u|^{1 / 2}}\right|=2 \varphi^{-1}(x)|t-x| . \tag{9}
\end{equation*}
$$

Further, for $g \in W(\varphi)$, we have $g(t)=g(x)+\int_{x}^{t} g^{\prime}(u) d u, t \in[0,1]$ and $x \in(0,1)$. Hence, by (9), Hölder's inequality, (4) and Lemma 1, d), we get

$$
\begin{aligned}
& \left|V_{n}(g ; x)-g(x)\right|=\left|V_{n}\left(\int_{x}^{t} g^{\prime}(u) d u ; x\right)\right| \\
& \quad \leq V_{n}\left(\left|\int_{x}^{t}\right| g^{\prime}(u)|d u| ; x\right) \leq\left\|\varphi g^{\prime}\right\| V_{n}\left(\left|\int_{x}^{t} \frac{d u}{\varphi(u)} d u\right| ; x\right)
\end{aligned}
$$

$$
\begin{align*}
& \leq 2 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\| V_{n}(|t-x| ; x) \leq 2 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\|\left(V_{n}\left((t-x)^{2} ; x\right)\right)^{1 / 2} \\
& =2 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\|\left(V_{n}\left(e_{2} ; x\right)-2 x V_{n}\left(e_{1} ; x\right)+x^{2} V_{n}\left(e_{0} ; x\right)\right)^{1 / 2} \\
& =2 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\|\left(x^{2}-2 x r_{n}(x)+x^{2}\right)^{1 / 2}=2 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\|\left(2 x\left(x-r_{n}(x)\right)\right)^{1 / 2} \\
& \leq 2 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\|\left(2 x \frac{2}{n}(1-x)\right)^{1 / 2}=\frac{4}{\sqrt{n}}\left\|\varphi g^{\prime}\right\| . \tag{10}
\end{align*}
$$

Due to (7), the estimation (10) is also valid for $x \in\{0,1\}$.
On the other hand, by (4), we obtain $\left|\left(V_{n} f\right)(x)\right| \leq \sum_{k=0}^{n} p_{n, k}\left(r_{n}(x)\right)\left|f\left(\frac{k}{n}\right)\right| \leq$ $\|f\| \sum_{k=0}^{n} p_{n, k}\left(r_{n}(x)\right) \leq\|f\|$, therefore

$$
\begin{equation*}
\left\|V_{n} f\right\| \leq\|f\| \tag{11}
\end{equation*}
$$

for all $f \in C[0,1]$.
Now, in view of (4), (10) and (11), we find that

$$
\begin{aligned}
\left|V_{n}(f ; x)-f(x)\right| & \leq\left|V_{n}(f-g ; x)\right|+\left|V_{n}(g ; x)-g(x)\right|+|g(x)-f(x)| \\
& \leq 2\|f-g\|+\frac{4}{\sqrt{n}}\left\|\varphi g^{\prime}\right\| \leq 4\left\{\|f-g\|+\frac{1}{\sqrt{n}}\left\|\varphi g^{\prime}\right\|\right\} .
\end{aligned}
$$

Taking the infimum on the right hand side over all $\mathrm{g} \in \mathrm{W}(\varphi)$, we obtain

$$
\left\|V_{n} f-f\right\| \leq 4 K_{1, \varphi}\left(f ; \frac{1}{\sqrt{n}}\right)
$$

Hence, by (6), we arrive at (8), which completes the proof.

## 3 Converse theorem

We begin with the following remark.
Remark 1 Due to (8), the condition $\omega_{\varphi}^{1}(f ; \delta) \leq C_{3} \delta^{\alpha}, \delta>0,0<\alpha<1$ implies that $\left\|V_{n} f-f\right\| \leq C_{4} n^{-\alpha / 2}, n \geq 1$.

In what follows, we establish the converse result of the statement given in Remark 1. To achieve this we need some lemmas.

Lemma 2 We have
a) $\left\|\varphi\left(\mathrm{V}_{\mathrm{n}} \mathrm{f}\right)^{\prime}\right\| \leq 8 \sqrt{\mathrm{n}}\|\mathrm{f}\|$ for $\mathrm{f} \in \mathrm{C}[0,1]$ and $\mathrm{n}=1,2, \ldots$;
b) $\left\|\varphi\left(\mathrm{V}_{\mathrm{n}} \mathrm{g}\right)^{\prime}\right\| \leq 32\left\|\varphi \mathrm{~g}^{\prime}\right\|$ for $\mathrm{g} \in \mathrm{W}(\varphi)$ and $\mathrm{n}=1,2, \ldots$

Proof. a) Let $x \in(0,1)$. By [1, p. 305, (2.1)], we have for the derivatives of $p_{n, k}$ that

$$
\begin{equation*}
p_{n, k}^{\prime}(x)=n\left[p_{n-1, k-1}(x)-p_{n-1, k}(x)\right]=\varphi^{-2}(x)(k-n x) p_{n, k}(x), \tag{12}
\end{equation*}
$$

where $\mathrm{k}=1,2, \ldots, \mathrm{n}$ and $\mathrm{p}_{\mathrm{n}-1,-1}(\mathrm{x})=\mathrm{p}_{\mathrm{n}-1, \mathrm{n}}(\mathrm{x})=0$. We distinguish two cases: $x \in\left(0, \frac{1}{n}\right]$. By (1), (12), Lemma 1, a) and (4), we get

$$
\begin{align*}
\left|\varphi(x)\left(V_{n} f\right)^{\prime}(x)\right| & =\varphi(x) r_{n}^{\prime}(x)\left|\sum_{k=0}^{n} p_{n, k}^{\prime}\left(r_{n}(x)\right) f\left(\frac{k}{n}\right)\right| \\
& =n \varphi(x) r_{n}^{\prime}(x)\left|\sum_{k=0}^{n}\left[p_{n-1, k-1}\left(r_{n}(x)\right)-p_{n-1, k}\left(r_{n}(x)\right)\right] f\left(\frac{k}{n}\right)\right| \\
& =n \varphi(x) r_{n}^{\prime}(x)\left|\sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left[f\left(\frac{k+1}{n}\right)-f\left(\frac{k}{n}\right)\right]\right| \\
& \leq n \varphi(x) r_{n}^{\prime}(x) \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left|f\left(\frac{k+1}{n}\right)-f\left(\frac{k}{n}\right)\right| \\
& \leq 2 n \varphi(x) r_{n}^{\prime}(x)\|f\| \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right) \leq 4 n \sqrt{x(1-x)}\|f\| \\
& \leq 4 \sqrt{n}\|f\| . \tag{13}
\end{align*}
$$

$x \in\left[\frac{1}{n}, 1\right)$. Using (1), Lemma 1, a), (12), Hölder's inequality, (4) and Lemma $\overline{1, \mathrm{c}), \mathrm{d}), \mathrm{e})}$, we get

$$
\begin{aligned}
& \left|\varphi(x)\left(V_{n} f\right)^{\prime}(x)\right|=\varphi(x) r_{n}^{\prime}(x)\left|\sum_{k=0}^{n} p_{n, k}^{\prime}\left(r_{n}(x)\right) f\left(\frac{k}{n}\right)\right| \\
& \quad \leq \quad 2 \varphi(x)\|f\| \sum_{k=0}^{n} \varphi^{-2}\left(r_{n}(x)\right)\left|k-n r_{n}(x)\right| p_{n, k}\left(r_{n}(x)\right) \\
& \quad \leq 2 \varphi(x) \varphi^{-2}\left(r_{n}(x)\right)\|f\|\left(\sum_{k=0}^{n}\left(k-n r_{n}(x)\right)^{2} p_{n, k}\left(r_{n}(x)\right)\right)^{1 / 2}
\end{aligned}
$$

$$
\begin{align*}
& =2 n \varphi(x) \varphi^{-2}\left(r_{n}(x)\right)\|f\|\left(V_{n}\left(e_{2} ; x\right)-2 r_{n}(x) V_{n}\left(e_{1} ; x\right)+r_{n}^{2}(x) V_{n}\left(e_{0} ; x\right)\right)^{1 / 2} \\
& =2 n \varphi(x) \varphi^{-2}\left(r_{n}(x)\right)\|f\|\left(x^{2}-2 r_{n}^{2}(x)+r_{n}^{2}(x)\right)^{1 / 2} \\
& =2 n \varphi(x) \varphi^{-2}\left(r_{n}(x)\right)\|f\|\left(x+r_{n}(x)\right)^{1 / 2}\left(x-r_{n}(x)\right)^{1 / 2} \\
& \leq 2 n \varphi(x) \varphi^{-2}\left(r_{n}(x)\right)\|f\| \sqrt{2 x} \sqrt{\frac{2}{n}(1-x)}=4 \sqrt{n} \frac{\varphi^{2}(x)}{\varphi^{2}\left(r_{n}(x)\right)}\|f\| \\
& =4 \sqrt{n} \frac{x}{r_{n}(x)} \frac{1-x}{1-r_{n}(x)}\|f\| \leq 4 \sqrt{n} \cdot 2 \cdot 1 \cdot\|f\|=8 \sqrt{n}\|f\| \tag{14}
\end{align*}
$$

Finally, by Lemma 1, a), we get $\varphi(0)\left(V_{n} f\right)^{\prime}(0)=\varphi(0) \sum_{k=0}^{n} p_{n, k}\left(r_{n}(0)\right) r_{n}^{\prime}(0) f\left(\frac{k}{n}\right)$ $=0$ and $\varphi(1)\left(V_{n} f\right)^{\prime}(1)=\varphi(1) \sum_{k=0}^{n} p_{n, k}\left(r_{n}(1)\right) r_{n}^{\prime}(1) f\left(\frac{k}{n}\right)=0$. Hence, due to (13) and (14), we obtain $\left\|\varphi\left(V_{n} f\right)^{\prime}\right\| \leq 8 \sqrt{n}\|f\|$, which was to be proved.
b) The proof is similar to the above. Let $x \in\left(0, \frac{1}{n}\right]$. Taking into account (1), (12), Lemma 1, a), (9), Hölder's inequality and (4), we get for $g \in W(\varphi)$ that

$$
\begin{aligned}
& \left|\left(V_{n} g\right)^{\prime}(x)\right|=n r_{n}^{\prime}(x)\left|\sum_{k=0}^{n}\left[p_{n-1, k-1}\left(r_{n}(x)\right)-p_{n-1, k}\left(r_{n}(x)\right)\right] g\left(\frac{k}{n}\right)\right| \\
& \quad=n r_{n}^{\prime}(x)\left|\sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left[g\left(\frac{k+1}{n}\right)-g\left(\frac{k}{n}\right)\right]\right| \\
& \leq 2 n \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left|g\left(\frac{k+1}{n}\right)-g\left(\frac{k}{n}\right)\right| \\
& \leq 2 n \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left\{\left|g\left(\frac{k+1}{n}\right)-g(x)\right|+\left|g\left(\frac{k}{n}\right)-g(x)\right|\right\} \\
& \leq 2 n \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left\{\left|\int_{x}^{\frac{k+1}{n}}\right| g^{\prime}(u)|d u|+\left|\int_{x}^{\frac{k}{n}}\right| g^{\prime}(u)|d u|\right\} \\
& \leq 2 n\left\|\varphi g^{\prime}\right\| \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left\{\left|\int_{x}^{\frac{k+1}{n}} \frac{d u}{\varphi(u)}\right|+\left|\int_{x}^{\frac{k}{n}} \frac{d u}{\varphi(u)}\right|\right\} \\
& \leq 4 n \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\| \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left\{\left|\frac{k+1}{n}-x\right|+\left|\frac{k}{n}-x\right|\right\}
\end{aligned}
$$

$$
\begin{align*}
\leq & 4 n \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\|\left\{\left(\sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left(\frac{k+1}{n}-x\right)^{2}\right)^{1 / 2}\right. \\
& \left.+\left(\sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left(\frac{k}{n}-x\right)^{2}\right)^{1 / 2}\right\} \tag{15}
\end{align*}
$$

Further, by (3), Lemma 1, c) and $x \in\left(0, \frac{1}{n}\right]$, we obtain

$$
\begin{align*}
& \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left(\frac{k}{n}-x\right)^{2}=\left(\frac{n-1}{n}\right)^{2} \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left(\frac{k}{n-1}\right)^{2} \\
&-2 x \frac{n-1}{n} \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right) \frac{k}{n-1}+x^{2} \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right) \\
&=\left(\frac{n-1}{n}\right)^{2}\left[r_{n}^{2}(x)+\frac{1}{n-1} r_{n}(x)\left(1-r_{n}(x)\right)\right]-2 x \frac{n-1}{n} r_{n}(x)+x^{2} \\
&= \frac{(n-1)(n-2)}{n^{2}} r_{n}^{2}(x)+\frac{n-1}{n}\left(\frac{1}{n}-2 x\right) r_{n}(x)+x^{2} \\
& \leq \frac{(n-1)(n-2)}{n^{2}} r_{n}^{2}(x)+\frac{n-1}{n}\left(\frac{1}{n}+2 x\right) r_{n}(x)+x^{2} \\
& \leq \frac{(n-1)(n-2)}{n^{2}} \frac{1}{n^{2}}+\frac{n-1}{n}\left(\frac{1}{n}+\frac{2}{n}\right) \frac{1}{n}+\frac{1}{n^{2}} \\
& \leq \frac{1}{n^{2}}+\frac{3}{n^{2}}+\frac{1}{n^{2}}=\frac{5}{n^{2}} . \tag{16}
\end{align*}
$$

Using the inequality $(a+b)^{2} \leq 2\left(a^{2}+b^{2}\right)$ with $a$ and $b$ real numbers, (16) and (3), we obtain

$$
\begin{align*}
& \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left(\frac{k+1}{n}-x\right)^{2} \\
& \quad \leq 2 \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right)\left(\frac{k}{n}-x\right)^{2}+2 \sum_{k=0}^{n-1} p_{n-1, k}\left(r_{n}(x)\right) \frac{1}{n^{2}} \\
& \quad \leq \frac{10}{n^{2}}+\frac{2}{n^{2}}=\frac{12}{n^{2}} \tag{17}
\end{align*}
$$

Combining (15), (16) and (17), we get

$$
\begin{equation*}
\left|\varphi(x)\left(V_{n} g\right)^{\prime}(x)\right| \leq 4 n\left\|\varphi g^{\prime}\right\|\left(\frac{\sqrt{12}}{n}+\frac{\sqrt{5}}{n}\right) \leq 23\left\|\varphi g^{\prime}\right\| \tag{18}
\end{equation*}
$$

Let $x \in\left[\frac{1}{n}, 1\right)$. For $g \in W(\varphi)$, by (1), (4), (12), Lemma 1, a), (9), Hölder's inequality, (4) and Lemma 1, c), d), e), we find that

$$
\begin{aligned}
\mid \varphi & (x)\left(V_{n} g\right)^{\prime}(x)|=\varphi(x)|\left(V_{n} g\right)^{\prime}(x)-g(x)\left(V_{n} e_{0}\right)^{\prime}(x) \mid \\
& =\varphi(x)\left|\sum_{k=0}^{n} p_{n, k}^{\prime}\left(r_{n}(x)\right) r_{n}^{\prime}(x)\left[g\left(\frac{k}{n}\right)-g(x)\right]\right| \\
& =\varphi(x) r_{n}^{\prime}(x)\left|\sum_{k=0}^{n} p_{n, k}^{\prime}\left(r_{n}(x)\right) \int_{x}^{\frac{k}{n}} g^{\prime}(u) d u\right| \\
& \leq 2 \varphi(x)\left|\sum_{k=0}^{n} \varphi^{-2}\left(r_{n}(x)\right)\left(k-n r_{n}(x)\right) p_{n, k}\left(r_{n}(x)\right) \int_{x}^{\frac{k}{n}} g^{\prime}(u) d u\right| \\
& \leq 2 \varphi(x) \varphi^{-2}\left(r_{n}(x)\right) \sum_{k=0}^{n}\left|k-n r_{n}(x)\right| p_{n, k}\left(r_{n}(x)\right)\left|\int_{x}^{\frac{k}{n}}\right| g^{\prime}(u)|d u| \\
& \leq 4 \varphi^{-2}\left(r_{n}(x)\right)\left\|\varphi g^{\prime}\right\| \sum_{k=0}^{n}\left|k-n r_{n}(x)\right|\left|\frac{k}{n}-x\right| p_{n, k}\left(r_{n}(x)\right) \\
& \leq 4 n \varphi^{-2}\left(r_{n}(x)\right)\left\|\varphi g^{\prime}\right\|\left(\sum_{k=0}^{n}\left(\frac{k}{n}-r_{n}(x)\right)^{2} p_{n, k}\left(r_{n}(x)\right)\right)^{1 / 2}
\end{aligned}
$$

$$
\times\left(\sum_{k=0}^{n}\left(\frac{k}{n}-x\right)^{2} p_{n, k}\left(r_{n}(x)\right)\right)^{1 / 2}
$$

$$
=4 n \varphi^{-2}\left(r_{n}(x)\right)\left\|\varphi g^{\prime}\right\|\left(V_{n}\left(e_{2} ; x\right)-2 r_{n}(x) V_{n}\left(e_{1} ; x\right)+r_{n}^{2}(x) V_{n}\left(e_{0} ; x\right)\right)^{1 / 2}
$$

$$
\times\left(V_{n}\left(e_{2} ; x\right)-2 x V_{n}\left(e_{1} ; x\right)+x^{2} V_{n}\left(e_{0} ; x\right)\right)^{1 / 2}
$$

$$
=4 n \varphi^{-2}\left(r_{n}(x)\right)\left\|\varphi g^{\prime}\right\|\left(x^{2}-r_{n}^{2}(x)\right)^{1 / 2}\left(2 x^{2}-2 x r_{n}(x)\right)^{1 / 2}
$$

$$
=4 \sqrt{2} n \varphi^{-2}\left(r_{n}(x)\right)\left\|\varphi g^{\prime}\right\|\left(x+r_{n}(x)\right)^{1 / 2} \sqrt{x}\left(x-r_{n}(x)\right)
$$

$$
\leq 4 \sqrt{2} n \varphi^{-2}\left(r_{n}(x)\right)\left\|\varphi g^{\prime}\right\| \sqrt{2 x} \sqrt{x} \frac{2}{n}(1-x)=16 \frac{\varphi^{2}(x)}{\varphi^{2}\left(r_{n}(x)\right)}\left\|\varphi g^{\prime}\right\|
$$

$$
\begin{equation*}
=16 \frac{x}{r_{n}(x)} \frac{1-x}{1-r_{n}(x)}\left\|\varphi g^{\prime}\right\| \leq 32\left\|\varphi g^{\prime}\right\| \tag{19}
\end{equation*}
$$

Finally, we have $\varphi(0)\left(\mathrm{V}_{\mathrm{n}} \mathrm{g}\right)^{\prime}(0)=0=\varphi(1)\left(\mathrm{V}_{\mathrm{n}} \mathrm{g}\right)^{\prime}(1)$. Hence, by (18) and (19), we obtain $\left\|\varphi\left(\mathrm{V}_{\mathrm{n}} \mathrm{g}\right)^{\prime}\right\| \leq 32\left\|\varphi \mathrm{~g}^{\prime}\right\|$, which completes the proof.

The next result is a weak-type version of the Berens-Lorentz lemma (see [1,
p. 312, Lemma 5.2]).

Lemma 3 Let $\phi:[0, a] \rightarrow[0, \infty)$ be an increasing function with $\phi(0)=0$ and $0<\alpha<1$. If $0<a \leq 1$, then the inequalities

$$
\begin{equation*}
\phi(\mathrm{a}) \leq \mathrm{C}_{5} \mathrm{a}^{\alpha} \tag{20}
\end{equation*}
$$

and

$$
\begin{equation*}
\phi(x) \leq C_{5}\left(y^{\alpha}+\frac{x}{y} \phi(y)\right), \quad 0 \leq x \leq y \leq a \tag{21}
\end{equation*}
$$

imply for some $\mathrm{C}_{6}=\mathrm{C}_{6}(\alpha)>0$ that

$$
\begin{equation*}
\phi(x) \leq C_{6} C_{5} x^{\alpha}, \quad 0 \leq x \leq a \tag{22}
\end{equation*}
$$

Proof. Following the proof of Lemma 5.2 in [1, p. 312], it is easy to prove our result taking into account the slight modification on $\alpha$. For completeness we give the proof.

For $0<q<1$, we define $x_{k}=q^{k} a, k=0,1,2, \ldots$ If we take $C \geq 1$, then (20) implies (22) for $x=x_{0}$. We prove (22) for all $x=x_{k}$ by induction. Let $\phi\left(x_{k}\right) \leq C_{5} x_{k}^{\alpha}$, then, by (21),

$$
\phi\left(x_{k+1}\right) \leq C_{5}\left(x_{k}^{\alpha}+q \phi\left(x_{k}\right)\right) \leq C_{5}\left(1+q C C_{5}\right) x_{k}^{\alpha} \leq C_{5} C_{x_{k+1}}^{\alpha}
$$

provided $1+\mathrm{qCC}_{5} \leq \mathrm{Cq}^{\alpha}$. To achieve this, we first take q so small that $\mathrm{q}^{\alpha}>\mathrm{C}_{5} \mathrm{q}$, because $0<\alpha<1$, and then C sufficiently large. After this, for any $0<x<a$, we select a $k$ with $x_{k+1} \leq x \leq x_{k}$ and get with $C_{6}:=\mathrm{Cq}^{-\alpha}$ the estimations $\phi(x) \leq \phi\left(x_{k}\right) \leq C_{5} x_{k}^{\alpha} \leq C_{6} C_{5} x^{\alpha}$.

In the next theorem we establish the converse result. We set $\mathrm{C}_{01}[0,1]=\{f \in$ $C[0,1]: f(0)=f(1)\}$.

Theorem 2 For $\mathrm{f} \in \mathrm{C}_{01}[0,1], 0<\alpha<1$ and $\mathrm{V}_{\mathrm{n}}$ defined by (1)-(2), the estimation

$$
\begin{equation*}
\left\|V_{n} f-f\right\| \leq C_{7} n^{-\alpha / 2}, \quad n=1,2, \ldots \tag{23}
\end{equation*}
$$

implies $\omega_{\varphi}^{1}(f ; \delta) \leq \mathrm{C}_{8} \delta^{\alpha}, 0<\delta \leq 1$.
Proof. The proof is based on Lemma 3 with $\phi(t)=\omega_{\varphi}^{1}(f ; t), t \in[0,1]$. For $f \in C_{01}[0,1]$, by Lemma $\left.1, b\right)$, we have

$$
\begin{aligned}
\left(V_{1} f\right)(x) & =p_{1,0}\left(r_{1}(x)\right) f(0)+p_{1,1}\left(r_{1}(x)\right) f(1)=\left(1-x^{2}\right) f(0)+x^{2} f(1) \\
& =f(0)+x^{2}(f(1)-f(0))=f(0)
\end{aligned}
$$

Therefore, by (5), $\omega_{\varphi}^{1}\left(\mathrm{f}-\mathrm{V}_{1} \mathrm{f} ; \mathrm{t}\right)=\omega_{\varphi}^{1}(\mathrm{f} ; \mathrm{t}), \mathrm{t}>0$. Hence, due to (5) and (23),

$$
\begin{equation*}
\omega_{\varphi}^{1}(f ; 1)=\omega_{\varphi}^{1}\left(f-V_{1} f ; 1\right) \leq 2\left\|f-V_{1} f\right\| \leq 2 C_{7} . \tag{24}
\end{equation*}
$$

Let $x \in[0,1]$ and $h>0$ such that $x \pm \frac{h}{2} \in[0,1]$, and let $\Delta_{h}^{1} f(x)=f\left(x+\frac{h}{2}\right)-$ $f\left(x-\frac{h}{2}\right)$. Then, by (23),

$$
\begin{align*}
& \left|\Delta_{h}^{1} f(x)\right| \leq\left|\Delta_{h}^{1}\left(f-V_{n} f\right)(x)\right|+\left|\Delta_{h}^{1}\left(V_{n} f\right)(x)\right| \\
& \quad \leq 2\left\|f-V_{n} f\right\|+\left|\Delta_{h}^{1}\left(V_{n} f\right)(x)\right| \leq 2 C_{7} n^{-\alpha / 2}+\left|\Delta_{h}^{1}\left(V_{n} f\right)(x)\right| . \tag{25}
\end{align*}
$$

Using (6), we can choose $g=g_{\delta} \in$ A.C.loc $[0,1]$ such that $\|f-g\| \leq C_{9} \omega_{\varphi}^{1}(f ; \delta)$ and $\left\|\varphi g^{\prime}\right\| \leq \mathrm{C}_{10} \delta^{-1} \omega_{\varphi}^{1}(\mathrm{f} ; \delta)$. Hence, in view of Lemma 2,

$$
\begin{aligned}
& \left|\left(V_{n} f\right)^{\prime}(x)\right| \leq\left|\left(V_{n}(f-g)\right)^{\prime}(x)\right|+\left|\left(V_{n} g\right)^{\prime}(x)\right| \\
& \leq 8 \sqrt{n} \varphi^{-1}(x)\|f-g\|+32 \varphi^{-1}(x)\left\|\varphi g^{\prime}\right\| \\
& \leq 8 \sqrt{n} C_{9} \varphi^{-1}(x) \omega_{\varphi}^{1}(f ; \delta)+32 C_{10} \delta^{-1} \varphi^{-1}(x) \omega_{\varphi}^{1}(f ; \delta) \\
& \quad \leq C_{11} \varphi^{-1}(x)\left(\sqrt{n}+\frac{1}{\delta}\right) \omega_{\varphi}^{1}(f ; \delta),
\end{aligned}
$$

where $C_{11}=8 C_{9}+32 C_{10}$. This implies that

$$
\left|\Delta_{h}^{1}\left(V_{n} f\right)(x)\right|=\left|\int_{x-\frac{h}{2}}^{x+\frac{h}{2}}\left(V_{n} f\right)^{\prime}(u) d u\right| \leq C_{11}\left(\sqrt{n}+\frac{1}{\delta}\right) \omega_{\varphi}^{1}(f ; \delta)\left|\int_{x-\frac{h}{2}}^{x+\frac{h}{2}} \frac{d u}{\varphi(u)}\right| .
$$

Because of $x \pm \frac{h}{2} \in[0,1]$, we have $x \in(0,1)$. Using (9), we obtain

$$
\begin{aligned}
\left|\int_{x-\frac{h}{2}}^{x+\frac{h}{2}} \frac{d u}{\varphi(u)}\right| & \leq\left|\int_{x}^{x+\frac{h}{2}} \frac{d u}{\varphi(u)}\right|+\left|\int_{x}^{x-\frac{h}{2}} \frac{d u}{\varphi(u)}\right| \\
& \leq 2 \varphi^{-1}(x) \frac{h}{2}+2 \varphi^{-1}(x) \frac{h}{2}=2 \varphi^{-1}(x) h .
\end{aligned}
$$

Hence, by (25), we get

$$
\begin{aligned}
\left|\Delta_{h}^{1} f(x)\right| & \leq 2 C_{7} n^{-\alpha / 2}+C_{11}\left(\sqrt{n}+\frac{1}{\delta}\right) \omega_{\varphi}^{1}(f ; \delta) 2 \varphi^{-1}(x) h \\
& \leq C_{12}\left\{n^{-\alpha / 2}+\left(h \sqrt{n}+\frac{h}{\delta}\right) \varphi^{-1}(x) \omega_{\varphi}^{1}(f ; \delta)\right\} .
\end{aligned}
$$

Replacing $h$ by $h \varphi(x)$ gives

$$
\left|\Delta_{h \varphi(x)}^{1} f(x)\right| \leq C_{12}\left\{n^{-\alpha / 2}+\left(h \sqrt{n}+\frac{h}{\delta}\right) \omega_{\varphi}^{1}(f ; \delta)\right\} .
$$

Now we choose $n \geq 1$ such that $\frac{1}{\sqrt{n}} \leq \delta \leq \frac{2}{\sqrt{n}}$, where $0<\delta \leq 1$. Then we find that

$$
\left|\Delta_{h \varphi(x)}^{1} f(x)\right| \leq C_{13}\left\{\delta^{\alpha}+\frac{h}{\delta} \omega_{\varphi}^{1}(f ; \delta)\right\}
$$

for all $x$ with $x \pm \frac{h}{2} \varphi(x) \in[0,1]$. Taking supremum over all $h$ with $0<h \leq t$, we obtain

$$
\begin{equation*}
\omega_{\varphi}^{1}(f ; t) \leq C_{13}\left\{\delta^{\alpha}+\frac{t}{\delta} \omega_{\varphi}^{1}(f ; \delta)\right\}, \quad 0<t \leq \delta . \tag{26}
\end{equation*}
$$

Now (24) and (26) yield the assertion of our theorem by Lemma 3.

## References

[1] R. A. DeVore, G. G. Lorentz, Constructive Approximation, Springer, Berlin, 1993.
[2] Z. Ditzian, V. Totik, Moduli of Smoothness, Springer, New York, 1987.
[3] J. P. King, Positive linear operators which preserve x ${ }^{2}$, Acta Math. Hungar., 99 (2003), 203-208.

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# Fixed point theorem for new type of auxiliary functions 

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#### Abstract

In this paper, we present some fixed point results satisfying generalized contractive condition with new auxiliary function in complete metric spaces. More precisely, the structure of the paper is the following. In the first section, we present some useful notions and results. The main aim of second section is to establish some new fixed point results in complete metric spaces. Finally, in the third section, we show the validity and superiority of our main results by suitable example. Also, as an application of our main result, some interesting corollaries have been included, which make our concepts and results effective. Our main result generalizes some well known existing results in the literature.


## 1 Introduction and preliminaries

The Banach contraction principle [10] is one of the revolutionary results of the fixed point theory, and it plays an imperative role to solve existence problems
in many branches of nonlinear analysis. Inspired from the impact of this natural idea to functional analysis, a number of researchers have been extended and generalized this principle for different kinds of contractions in various spaces.

Let us denote:
$\Psi_{1}=\left\{\psi_{1}:[0, \infty) \rightarrow[0, \infty)\right.$ is a continuous and non-decreasing function such that $\psi_{1}(t)=0$ if and only if $\left.t=0.\right\}$ (Altering distance function)
$\Psi_{2}=\left\{\psi_{2}:[0, \infty) \rightarrow[0, \infty)\right.$ is a continuous function such that $\psi_{2}(0) \geq 0$ and $\left.\psi_{2}(\mathrm{t})>0, \mathrm{t}>0.\right\}$ (Ultra-altering distance function)
$\Psi_{3}=\{\varphi:[0, \infty) \rightarrow[0, \infty)$ is a Lebesgue-integrable function, summable on each compact subset of $\mathrm{R}^{+}$, non-negative, and such that for each $\epsilon>0$, $\left.\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}>0.\right\}$

In 2002, Branciari [3] introduced one of the genuine contraction, known as integral type contraction, as an analogue of Banach contraction principle [10].

Theorem $1[3] \operatorname{Let}(\mathrm{E}, \mathrm{d})$ be a complete metric space, $\mathrm{k} \in(0,1)$, and $A: E \rightarrow$ E is such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$

$$
\begin{equation*}
\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt} \leq \mathrm{k} \int_{0}^{\mathrm{d}(x, y)} \varphi(\mathrm{t}) \mathrm{dt} \tag{1}
\end{equation*}
$$

where $\varphi \in \Psi_{3}$. Then A has a unique fixed point of $z \in E$.
Rhoades [5], in 2003, gave an extension of the result of Branciari [3] and proved following theorems.

Theorem 2 [5] Let (E, d) be a complete metric space and $A: E \rightarrow E$ be a mapping such that,

$$
\int_{0}^{\mathrm{d}(A x, A y)} \varphi(t) d t \leq \beta \int_{0}^{M(x, y)} \varphi(t) d t
$$

where

$$
\begin{equation*}
M(x, y)=\max \left\{d(x, y), d(x, A x), d(y, A y), \frac{d(x, A y)+d(y, A x)}{2}\right\} \tag{2}
\end{equation*}
$$

for all $x, y \in E, \beta \in[0,1)$ and $\varphi \in \Psi_{3}$. Then $A$ has a unique fixed point $z \in E$.
Theorem 3 [5] Let us consider a complete metric space ( $\mathrm{E}, \mathrm{d}$ ) and $\mathrm{A}: \mathrm{E} \rightarrow \mathrm{E}$ is a mapping such that,

$$
\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt} \leq \beta \int_{0}^{\mathrm{N}(x, y)} \varphi(\mathrm{t}) \mathrm{dt}
$$

where

$$
\begin{equation*}
\mathrm{N}(\mathrm{x}, \mathrm{y})=\max \{\mathrm{d}(\mathrm{x}, \mathrm{y}), \mathrm{d}(\mathrm{x}, A x), \mathrm{d}(\mathrm{y}, A \mathrm{y}), \mathrm{d}(\mathrm{x}, A \mathrm{y}), \mathrm{d}(\mathrm{y}, A x)\} \tag{3}
\end{equation*}
$$

for each $x, y \in E, \beta \in[0,1)$ and $\varphi \in \Psi_{3}$. Then there exist a unique fixed point $z \in E$ such that $A z=z$.

In 2010, Babu and Alemayehu [7] proved following theorem in complete metric spaces by using generalized $\phi-$ weak contraction.

Theorem $4[7]$ Let us consider a complete metric space ( $\mathrm{E}, \mathrm{d}$ ) and $\mathrm{A}: \mathrm{E} \rightarrow \mathrm{E}$ is such that for all $\mathrm{x}, \mathrm{y} \in \mathrm{E}$ it satisfies

$$
\mathrm{d}(A x, A y) \leq M(x, y)-\phi(M(x, y))
$$

where

$$
\begin{equation*}
M(x, y)=\max \left\{d(x, y), d(x, A x), d(y, A y), \frac{d(x, A y)+d(y, A x)}{2}\right\} \tag{4}
\end{equation*}
$$

and $\phi:[0,+\infty) \rightarrow[0,+\infty)$ such that $\phi(\mathrm{t})=0$ if and only if $\mathrm{t}=0$. Then there is a unique fixed point of A in E .

In 2011, Samet and Yazidi [6] gave an extension of the result of Dass and Gupta [4] in the sense of Branciari integral type contraction, as follows

Theorem 5 [6] Let (E, d) be a complete metric space and A be a self-map of E such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt} \leq \alpha \int_{0}^{\mathrm{M}(\mathrm{x}, \mathrm{y})} \varphi(\mathrm{t}) \mathrm{dt}+\beta \int_{0}^{\mathrm{d}(x, y)} \varphi(\mathrm{t}) \mathrm{dt}
$$

and

$$
M(x, y)=\frac{d(y, A y)[1+d(x, A x)]}{[1+d(x, y)]}
$$

where $\alpha, \beta>0$ are constants such that $\alpha+\beta<1$ and $\varphi \in \Psi_{3}$.
Then $\mathcal{A}$ admits a unique fixed point $z \in E$ such that for each $z \in E, A^{n} x \rightarrow z$ as $\mathrm{n} \rightarrow \infty$.

In 2011, Gupta and Mani [14] proved a common fixed point theorem for two weakly compatible mappings using control functions $\psi_{1}$ and $\psi_{2}$ satisfying a contractive condition of integral type.

Theorem 6 Let A and B be self compatible maps of a complete metric space ( $\mathrm{E}, \mathrm{d}$ ) satisfying the following conditions:
(i). $A(E) \subset B(E)$
(ii). $\psi_{1}\left(\int_{0}^{d(A x, A y)} \varphi(t) d t\right) \leq \psi_{1}\left(\int_{0}^{d(B x, B y)} \varphi(t) d t\right)-\psi_{2}\left(\int_{0}^{d(B x, B y)} \varphi(t) d t\right)$,
where $\psi_{1} \in \Psi_{1}, \psi_{2} \in \Psi_{2}, \varphi \in \Psi_{3}$. Then there exist a unique common fixed point of A and B in E .

In 2013, Gupta and Mani [16] proved another generalization of the result of Branciari [3] using real valued function.

Theorem 7 [16] Let $A$ be a self map on complete metric space ( $\mathrm{E}, \mathrm{d}$ ) such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$

$$
\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt} \leq \gamma(\mathrm{d}(\mathrm{x}, \mathrm{y})) \int_{0}^{\mathrm{m}(x, y)} \varphi(\mathrm{t}) \mathrm{dt}
$$

and

$$
m(x, y)=\max \left\{\frac{d(x, A x) d(y, A y)}{d(x, y)}, d(x, y)\right\}
$$

where $\varphi \in \Psi_{3}$ and $\gamma: \mathrm{R}^{+} \rightarrow[0,1)$ is a function with

$$
\lim _{\delta \rightarrow \mathrm{t}} \sup \gamma(\delta)<1, \quad \forall \quad \mathrm{t}>0
$$

Then A has a unique fixed point in E.
Some other results in complete metric spaces satisfying integral type contractions are mentioned in $[8,9,11,12,13,17,18]$

In 2014-15, Ansari [1] introduced the notion of C -class function as a major generalization of Banach contraction principle. Currently this finding is one of the most attractive research topics in fixed point theory. Some other special cases of C-class functions can be found in [2]. Ansari [1] gave the following definitions and examples.

Definition 1 [1] A mapping $\mathrm{F}:[0, \infty)^{2} \rightarrow \mathbb{R}$ is called C -class function if it is continuous and satisfies following axioms:

1. $\mathrm{F}(\mathrm{r}, \mathrm{t}) \leq \mathrm{r}$ for all $\mathrm{r}, \mathrm{t} \in[0, \infty)$;
2. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}$ implies that either $\mathrm{r}=0$ or $\mathrm{t}=0$;

Let us denote $\mathcal{C}$ the family of $\mathbf{C}$-class functions.
Remark 1 Clearly, for some F we have $\mathrm{F}(0,0)=0$.
Example 1 [1] The following functions $F:[0, \infty)^{2} \rightarrow \mathbb{R}$ are elements of $\mathcal{C}$, for all $r, t \in[0, \infty)$ :

1. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}-\mathrm{t}, \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{t}=0$;
2. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{mr}, 0<\mathrm{m}<1, \mathrm{~F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{r}=0$;
3. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\frac{\mathrm{r}}{(1+\mathrm{t})^{h}} ; \mathrm{h} \in(0, \infty), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{r}=0$ or $\mathrm{t}=0$;
4. $F(r, t)=\log \left(t+a^{r}\right) /(1+t), a>1, F(r, t)=r \Rightarrow r=0$ or $t=0$;
5. $F(r, t)=\ln \left(1+a^{r}\right) / 2, a>e, F(r, t)=r \Rightarrow r=0$;
6. $\mathrm{F}(\mathrm{r}, \mathrm{t})=(\mathrm{r}+\mathrm{l})^{\left(1 /(1+\mathrm{t})^{\mathrm{p}}\right)}-\mathrm{l}, \mathrm{l}>1, \mathrm{p} \in(0, \infty), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{t}=0$;
7. $F(r, t)=r \log _{t+a} a, a>1, F(r, t)=r \Rightarrow r=0$ or $t=0$;
8. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}-\left(\frac{1+\mathrm{r}}{2+\mathrm{r}}\right)\left(\frac{\mathrm{t}}{1+\mathrm{t}}\right), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{t}=0$;
9. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \beta(\mathrm{r}), \beta:[0, \infty) \rightarrow[0,1), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{r}=0$;
10. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}-\frac{\mathrm{t}}{\mathrm{k}+\mathrm{t}}, \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{t}=0$;
11. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}-\varphi(\mathrm{r}), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{r}=0$, here $\varphi:[0, \infty) \rightarrow[0, \infty)$ is a continuous function such that $\varphi(\mathrm{t})=0 \Leftrightarrow \mathrm{t}=0$;
12. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{rh}(\mathrm{r}, \mathrm{t}), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{r}=0$, here $\mathrm{h}:[0, \infty) \times[0, \infty) \rightarrow[0, \infty)$ is a continuous function such that $\mathrm{h}(\mathrm{r}, \mathrm{t})<1$ for all $\mathrm{t}, \mathrm{s}>0$;
13. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}-\left(\frac{2+\mathrm{t}}{1+\mathrm{t}}\right) \mathrm{t}, \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{t}=0$.
14. $F(r, t)=\sqrt[n]{\ln \left(1+r^{n}\right)}, F(r, t)=r \Rightarrow r=0$.
15. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\phi(\mathrm{r}), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r} \Rightarrow \mathrm{r}=0$, here $\phi:[0, \infty) \rightarrow[0, \infty)$ is a upper semi-continuous function such that $\phi(0)=0$, and $\phi(\mathrm{t})<\mathrm{t}$ for $\mathrm{t}>0$,
16. $\mathrm{F}(\mathrm{r}, \mathrm{t})=\frac{\mathrm{r}}{(1+\mathrm{r})^{s}} ; \mathrm{s} \in(0, \infty), \mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{r}$ implies $\mathrm{r}=0$.

Remark 2 We assume that is F increasing with respect to the first variable and decreasing with respect to the second variable.

The aim of this contribution is to investigate some fixed point results using the concept of $C$-class function and control functions in the set up of complete metric spaces satisfying a generalized weak contraction. Our result mainly generalized the result of Rhoades [5] and Gupta and Mani [15].

## 2 Main result- fixed point results with auxiliary functions

The main result of this paper is the following theorem.
Theorem 8 Let $(\mathrm{E}, \mathrm{d})$ be a complete metric space and $\mathcal{A}: \mathrm{E} \rightarrow \mathrm{E}$ be a mapping such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\begin{equation*}
\psi_{1}\left(\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt}\right) \leq \mathrm{F}\left(\psi_{1}\left(\int_{0}^{M(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right), \psi_{2}\left(\int_{0}^{M(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right)\right) \tag{5}
\end{equation*}
$$

where F is a C -class function, $\psi_{1} \in \Psi_{1}, \Psi_{2} \in \Psi_{2}, \varphi \in \Psi_{3}$ and

$$
\begin{equation*}
M(x, y)=\max \left\{d(x, y), d(x, A x), d(y, A y), \frac{d(x, A y)+d(y, A x)}{2}\right\} \tag{6}
\end{equation*}
$$

Then A has a unique fixed point.
Proof. Let $x_{0} \in E$ be an arbitrary point. Choose a point $x_{1}$ in $E$ such that $x_{1}=A x_{0}$. In general, choose $x_{n+1}$ such that $x_{n+1}=A x_{n}$ for $n=0,1,2 \cdots$.
Suppose that $x_{n} \neq x_{n+1}$ for each integer $n>1$, then from (5)

$$
\psi_{1}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right) \leq F\left\{\begin{array}{l}
\psi_{1}\left(\int_{0}^{M\left(x_{n-1}, x_{n}\right)} \varphi(t) d t\right)  \tag{7}\\
\psi_{2}\left(\int_{0}^{M\left(x_{n-1}, x_{n}\right)} \varphi(t) d t\right)
\end{array}\right\}
$$

where from (6),

$$
M\left(x_{n-1}, x_{n}\right)=\max \left\{\begin{array}{c}
d\left(x_{n-1}, x_{n}\right), d\left(x_{n-1}, x_{n}\right), d\left(x_{n}, x_{n+1}\right), \\
\frac{d\left(x_{n-1}, x_{n+1}\right)+d\left(x_{n}, x_{n}\right)}{2}
\end{array}\right\}
$$

$$
\begin{align*}
& =\max \left\{d\left(x_{n-1}, x_{n}\right), d\left(x_{n}, x_{n+1}\right), \frac{d\left(x_{n-1}, x_{n+1}\right)}{2}\right\} \\
& =\max \left\{d\left(x_{n-1}, x_{n}\right), d\left(x_{n}, x_{n+1}\right)\right\} \tag{8}
\end{align*}
$$

If $d\left(x_{n}, x_{n+1}\right) \geq d\left(x_{n-1}, x_{n}\right)$ for some $n$, then on combining equation (7) and (8), we get

$$
\psi_{1}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right) \leq F\left\{\begin{array}{l}
\psi_{1}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right)  \tag{9}\\
\psi_{2}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right)
\end{array}\right\}
$$

Thus by definition of $F \in \mathcal{C}$, we get

$$
\text { either } \psi_{1}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right)=0 \text { or } \psi_{2}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right)=0
$$

From definition of $\psi_{1}$ and $\psi_{2}$, it is possible only if

$$
\int_{0}^{\mathrm{d}\left(x_{n}, x_{n+1}\right)} \varphi(\mathrm{t}) \mathrm{dt}=0
$$

This is a contradiction to our hypothesis. Thus $d\left(x_{n}, x_{n+1}\right)<d\left(x_{n-1}, x_{n}\right)$, this implies

$$
\begin{aligned}
\psi_{1}\left(\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t\right) & \leq F\left(\psi_{1} \int_{0}^{d\left(x_{n-1}, x_{n}\right)} \varphi(t) d t, \psi_{1} \int_{0}^{d\left(x_{n-1}, x_{n}\right)} \varphi(t) d t\right) \\
& \leq \psi_{1} \int_{0}^{d\left(x_{n-1}, x_{n}\right)} \varphi(t) d t
\end{aligned}
$$

Since $\psi_{1}$ is continuous and non-decreasing, therefore

$$
\int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t \leq \int_{0}^{d\left(x_{n-1}, x_{n}\right)} \varphi(t) d t
$$

thus $\left\{\int_{0}^{\left(d\left(x_{n}, x_{n+1}\right)\right.} \varphi(t) d t\right\}$ is monotone decreasing and lower bounded sequence. Therefore there exist $r \geq 0$ such that

$$
\begin{equation*}
\lim _{n \rightarrow \infty} \int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t=r \tag{10}
\end{equation*}
$$

Suppose that $\mathrm{r}>0$. Taking limit as $\mathrm{n} \rightarrow \infty$ on both sides of eq. (9) and using eq. (10), we get

$$
\psi_{1}(r) \leq F\left(\psi_{1}(r), \psi_{2}(r)\right),
$$

implies from definition of $\mathrm{F} \in \mathcal{C}$ that

$$
\text { either } \psi_{1}(r)=0 \text { or } \psi_{2}(r)=0
$$

Consequently, by definition of $\psi_{1}$ and $\psi_{2}$, we get $r=0$.
Hence from eq. (10), we obtain

$$
\begin{equation*}
\lim _{n \rightarrow \infty} \int_{0}^{d\left(x_{n}, x_{n+1}\right)} \varphi(t) d t=0 \tag{11}
\end{equation*}
$$

implies

$$
\begin{equation*}
\lim _{n \rightarrow \infty} d\left(x_{n}, x_{n+1}\right)=0 \tag{12}
\end{equation*}
$$

Next we prove that $\left\{x_{n}\right\}$ is a Cauchy sequence. Suppose it is not. Therefore for an $\epsilon>0$, there exists two sub-sequences $\left\{x_{\mathfrak{m}(p)}\right\}$ and $\left\{x_{\mathfrak{n}(p)}\right\}$ of $\left\{x_{n}\right\}$ with $\mathfrak{m}(\mathfrak{p})<\mathfrak{n}(\mathfrak{p})<\mathfrak{m}(\mathfrak{p}+1)$ such that

$$
\begin{equation*}
\mathrm{d}\left(x_{\mathfrak{m}(\mathfrak{p})}, x_{\mathfrak{n}(\mathfrak{p})}\right) \geq \epsilon, \quad \mathrm{d}\left(x_{\mathrm{m}(\mathfrak{p})}, x_{\mathrm{n}(\mathfrak{p})-1}\right)<\epsilon . \tag{13}
\end{equation*}
$$

Consider

$$
\begin{align*}
\psi_{1}\left(\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}\right) & \leq \psi_{1}\left(\int_{0}^{\mathrm{d}\left(x_{m(p)}, x_{n(p)}\right)} \varphi(\mathrm{t}) \mathrm{dt}\right) \\
& \leq \mathrm{F}\left\{\begin{array}{l}
\psi_{1}\left(\int_{0}^{M\left(x_{m(p)-1}, x_{n(p)-1}\right)} \varphi(\mathrm{t}) \mathrm{dt}\right), \\
\psi_{2}\left(\int_{0}^{M\left(x_{m(p)-1}, x_{n(p)-1}\right)} \varphi(\mathrm{t}) \mathrm{dt}\right)
\end{array}\right\} . \tag{14}
\end{align*}
$$

Using (6)

$$
\begin{align*}
& M\left(x_{\mathfrak{m}(\mathfrak{p})-1}, x_{\mathfrak{n}(\mathfrak{p})-1}\right)=\max \left\{\begin{array}{c}
d\left(x_{\mathfrak{m}(\mathfrak{p})-1}, x_{\mathfrak{n}(\mathfrak{p})-1}\right), d\left(x_{\mathfrak{m}(\mathfrak{p})-1}, x_{\mathfrak{m}(\mathfrak{p})}\right), \\
d\left(x_{\mathfrak{n}(\mathfrak{p})-1}, x_{\mathfrak{n}(p)}\right), x_{1} \\
\frac{d\left(x_{\mathfrak{m}(\mathfrak{p})-1}, x_{\mathfrak{n}(\mathfrak{p})}\right)+\mathrm{d}\left(x_{\mathfrak{n}(\mathfrak{p})-1}, x_{\mathfrak{m}(\mathfrak{p})}\right)}{2}
\end{array}\right\} \\
& =\max \left\{\begin{array}{c}
d\left(x_{\mathfrak{m}(\mathfrak{p})-1}, x_{\mathfrak{n}(\mathfrak{p})-1}\right), d\left(x_{\mathfrak{m}(\mathfrak{p})-1}, x_{\mathfrak{m}(\mathfrak{p})}\right), \\
d\left(x_{\mathfrak{n}(\mathfrak{p})-1}, x_{\mathfrak{n}(\mathfrak{p})}\right), z(\mathfrak{m}, \mathfrak{n})
\end{array}\right\}, \tag{15}
\end{align*}
$$

where,

$$
\begin{equation*}
z(m, n)=\frac{d\left(x_{\mathfrak{m}(p)-1, x_{n(p)}}\right)+d\left(x_{n(p)-1, x_{\mathfrak{m}(p)}}\right)}{2} \tag{16}
\end{equation*}
$$

Thus

$$
\begin{aligned}
& \int_{0}^{M\left(x_{m(p)-1, x_{n}(p)-1}\right)} \varphi(t) d t
\end{aligned}
$$

Using (13) and triangle inequality, we get

$$
\begin{aligned}
\mathrm{d}\left(x_{\mathfrak{m}(p)-1}, x_{\mathfrak{n}(p)-1}\right) & \leq \mathrm{d}\left(x_{\mathfrak{m}(p)-1}, x_{\mathfrak{m}(p)}\right)+\mathrm{d}\left(x_{\mathfrak{m}(\mathfrak{p})}, x_{\mathfrak{n}(\mathfrak{p})-1}\right) \\
& <\mathrm{d}\left(x_{\mathfrak{m}(p)-1}, x_{\mathfrak{m}(p)}\right)+\epsilon
\end{aligned}
$$

Therefore,

$$
\begin{equation*}
\lim _{p \rightarrow \infty} \int_{0}^{d\left(x_{m}(\mathfrak{p})-1, x_{n(p)-1}\right)} \varphi(t) d t \leq \int_{0}^{\epsilon} \varphi(t) d t \tag{18}
\end{equation*}
$$

Also,

$$
\begin{aligned}
z(m, n) & =\frac{d\left(x_{\mathfrak{m}(p)-1}, x_{\mathfrak{n}(\mathfrak{p})}\right)+d\left(x_{\mathfrak{n}(p)-1}, x_{\mathfrak{m}(p)}\right)}{2} \\
& \leq \frac{d\left(x_{\mathfrak{m}(p)-1}, x_{\mathfrak{m}(p)}\right)+2 d\left(x_{\mathfrak{m}(\mathfrak{p})}, x_{\mathfrak{n}(p)-1}\right)+d\left(x_{\mathfrak{n}(p)-1}, x_{\mathfrak{n}(\mathfrak{p})}\right)}{2} \\
& \leq \frac{d\left(x_{\mathfrak{m}(p)-1}, x_{\mathfrak{m}(p)}\right)+d\left(x_{\mathfrak{n}(p)-1}, x_{\mathfrak{n}(p))}\right.}{2}+\epsilon
\end{aligned}
$$

Taking limit as $p \rightarrow \infty$ and using (12), we get

$$
\begin{equation*}
\lim _{p \rightarrow \infty} \int_{0}^{z(m, n)} \varphi(\mathrm{t}) \mathrm{dt} \leq \int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt} \tag{19}
\end{equation*}
$$

Taking limit as $p \rightarrow \infty$ in equality (14) and using equations (15), (16), (17), (18) and (19) all together in (14), we get

$$
\psi_{1}\left(\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}\right) \leq \mathrm{F}\left(\psi_{1}\left(\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}\right), \psi_{2}\left(\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}\right)\right) .
$$

Again from definition of $F \in \mathcal{C}$ we get,

$$
\text { either } \psi_{1}\left(\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}\right)=0 \quad \text { or } \psi_{2}\left(\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}\right)=0
$$

It is possible only if, $\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}=0$. This is a contradiction to our hypothesis. Therefore $\left\{x_{n}\right\}$ is a Cauchy sequence, call the limit $\alpha$ such that

$$
\begin{equation*}
\lim _{n \rightarrow \infty} A x_{n-1}=\alpha \tag{20}
\end{equation*}
$$

Next we claim that $\alpha$ is the fixed point of map $A$.
That is $A \alpha=\alpha$, suppose it is not. Then $d(A \alpha, \alpha)>0$.
Let $\delta=\mathrm{d}(A \alpha, \alpha)$.
Consider,

$$
\begin{align*}
\psi_{1}\left(\int_{0}^{\delta} \varphi(\mathrm{t}) \mathrm{dt}\right) & =\psi_{1}\left(\int_{0}^{\mathrm{d}(\mathrm{~A} \alpha, \alpha)} \varphi(\mathrm{t}) \mathrm{dt}\right) \\
& \leq \mathrm{F}\left\{\begin{array}{l}
\psi_{1}\left(\int_{0}^{\mathrm{M}\left(\alpha, x_{n}\right)} \varphi(\mathrm{t}) \mathrm{dt}\right) \\
\psi_{2}\left(\int_{0}^{\mathrm{M}\left(\alpha, x_{n}\right)} \varphi(\mathrm{t}) \mathrm{dt}\right)
\end{array}\right\} \tag{21}
\end{align*}
$$

where,

$$
M\left(\alpha, x_{n}\right)=\max \left\{\begin{array}{c}
d\left(\alpha, x_{n}\right), d(\alpha, A \alpha), d\left(x_{n}, x_{n+1}\right)  \tag{22}\\
\frac{d\left(\alpha, x_{n+1}\right)+d\left(x_{n}, A \alpha\right)}{2}
\end{array}\right\} .
$$

Since,

$$
\begin{equation*}
\lim _{n \rightarrow \infty} d\left(\alpha, x_{n}\right)=\lim _{n \rightarrow \infty} d\left(x_{n}, x_{n+1}\right)=\lim _{n \rightarrow \infty} d\left(\alpha, x_{n+1}\right)=0 \tag{23}
\end{equation*}
$$

Taking $\lim _{\mathfrak{n} \rightarrow \infty}$ in (21) and by using (20), (22), (23), we get

$$
\psi_{1}\left(\int_{0}^{\delta} \varphi(\mathrm{t}) \mathrm{dt}\right) \leq \mathrm{F}\left\{\begin{array}{l}
\psi_{1}\left(\int_{0}^{\max \left\{\mathrm{d}(\alpha, A \alpha), \frac{\mathrm{d}(\alpha, A \alpha)}{2}\right\}} \varphi(\mathrm{t}) \mathrm{dt}\right) \\
\psi_{2}\left(\int_{0}^{\max \left\{\mathrm{d}(\alpha, A \alpha), \frac{\mathrm{d}(\alpha, A \alpha)}{2}\right\}} \varphi(\mathrm{t}) \mathrm{dt}\right)
\end{array}\right\}
$$

$$
\begin{equation*}
\leq \mathrm{F}\left\{\psi_{1}\left(\int_{0}^{\delta} \varphi(\mathrm{t}) \mathrm{dt}\right), \psi_{2}\left(\int_{0}^{\delta} \varphi(\mathrm{t}) \mathrm{dt}\right)\right\} \tag{24}
\end{equation*}
$$

Thus we obtain,

$$
\text { either } \psi_{1}\left(\int_{0}^{\delta} \varphi(t) d t\right)=0 \text { or } \psi_{2}\left(\int_{0}^{\delta} \varphi(t) d t\right)=0
$$

that is $\int_{0}^{\delta} \varphi(t) d t=0$. Hence $\delta=0$.
This implies $\mathrm{d}(A \alpha, \alpha)=0$. Therefore $\alpha$ is the fixed point of map $A$. Uniqueness of the fixed point can be easily obtain by using above inequality (21), (22), (24). This proves the main result.

## 3 Applications and example

Next we give several corollaries, as a application of our main result, in the underlying spaces. Some of them are novel in literature

If we take $\psi_{1}(t)=t$ in Theorem 8 , we get a new result.
Corollary 1 Let ( $\mathrm{E}, \mathrm{d}$ ) be a complete metric space and $\mathrm{A}: \mathrm{E} \rightarrow \mathrm{E}$ be a mapping such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt} \leq \mathrm{F}\left(\int_{0}^{\mathrm{M}(x, y)} \varphi(\mathrm{t}) \mathrm{dt}, \psi_{2}\left(\int_{0}^{\mathrm{M}(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right)\right)
$$

for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$, where $\mathrm{M}(\mathrm{x}, \mathrm{y})$ is given in (6), F is a C -class function, $\psi_{2} \in \Psi_{2}, \varphi \in \Psi_{3}$.
Then A has a unique fixed point.
If we take $F(r, t)=\frac{r}{(1+t)^{s}}$ and assume $s=1$ in Theorem 8, we find a very interesting novel result.

Corollary 2 Let ( $\mathrm{E}, \mathrm{d}$ ) be a complete metric space and $\mathrm{A}: \mathrm{E} \rightarrow \mathrm{E}$ be a mapping such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\psi_{1}\left(\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt}\right) \leq \frac{\psi_{1}\left(\int_{0}^{M(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right)}{1+\psi_{2}\left(\int_{0}^{M(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right)}
$$

where $M(x, y)$ is given in (6), $\psi_{1} \in \Psi_{1}, \psi_{2} \in \Psi_{2}, \varphi \in \Psi_{3}$.
Then A has a unique fixed point.

If we take $F(r, t)=\lambda r$ for $0<\lambda<1$ and in Theorem 8 , then we have following corollary.

Corollary 3 Let ( $\mathrm{E}, \mathrm{d}$ ) be a complete metric space and $\mathrm{A}: \mathrm{E} \rightarrow \mathrm{E}$ be a mapping such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\psi_{1}\left(\int_{0}^{d(A x, A y)} \varphi(t) d t\right) \leq \lambda \psi_{1}\left(\int_{0}^{M(x, y)} \varphi(t) d t\right)
$$

where $M(x, y)$ is given in (6), $\psi_{1} \in \Psi_{1}, \varphi \in \Psi_{3}$.
Then A has a unique fixed point.
If we assume that $\psi_{1}(\mathrm{t})=\mathrm{t}$ in Corollary 3 then we obtain the result of Rhoades [5] (see Theorem 2 of [5]).

Corollary 4 Let ( $\mathrm{E}, \mathrm{d}$ ) be a complete metric space and $\mathrm{A}: \mathrm{E} \rightarrow \mathrm{E}$ be a mapping such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\int_{0}^{d(A x, A y)} \varphi(t) d t \leq \lambda \int_{0}^{M(x, y)} \varphi(t) d t
$$

where $M(x, y)$ is given in (6), $\varphi \in \Psi_{3}$.
Then A has a unique fixed point.
If we take $F(r, t)=r-t$ in Theorem 8 , then we obtain the result of Gupta and Mani [15].

Corollary 5 Let $(\mathrm{E}, \mathrm{d})$ be a complete metric space and $\mathcal{A}: \mathrm{E} \rightarrow \mathrm{E}$ be a mapping such that for each $\mathrm{x}, \mathrm{y} \in \mathrm{E}$,

$$
\psi_{1}\left(\int_{0}^{\mathrm{d}(A x, A y)} \varphi(\mathrm{t}) \mathrm{dt}\right) \leq \psi_{1}\left(\int_{0}^{M(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right)-\psi_{2}\left(\int_{0}^{M(x, y)} \varphi(\mathrm{t}) \mathrm{dt}\right)
$$

where $\mathcal{M}(x, y)$ is given in (6), $\psi_{1} \in \Psi_{1}, \psi_{2} \in \Psi_{2}, \varphi \in \Psi_{3}$.
Then A has a unique fixed point.
Remark 3 It should be noted that in [15], authors have considered an extra condition on $\psi_{2}$. But from above corollaries it is clear that we can deduce the same result without that extra assumption. Also, the result obtained in [5] and [15] are an element of $\mathcal{C}$ - class function as shown in Corollary 4 and Corollary 5. So the main result of this paper is more generalized than the other previously proved results.

Now, we gave a non trivial example to justify the importance of our result.
Example 2 Let $\mathrm{E}=\mathrm{N}-\{\infty\}$ and d is usual metric on E . Define a self maps A on E such that

$$
A x=\frac{x}{3}, \quad \forall \quad x \in E
$$

Define a function $\mathrm{F}:[0, \infty)^{2} \rightarrow \mathbb{R}$ as

$$
\mathrm{F}(\mathrm{r}, \mathrm{t})=\mathrm{mr}, \quad \forall 0<\mathrm{m}=\frac{2}{3}<1
$$

Then clearly, F is a C -class function.
Let us define $\psi_{1}, \varphi:[0,+\infty) \rightarrow[0,+\infty)$ as

$$
\psi_{1}(\mathrm{t})=\mathrm{t}, \quad \varphi(\mathrm{t})=\frac{\mathrm{t}}{2}, \quad \forall \quad \mathrm{t} \in[0,+\infty)
$$

then for each $\epsilon>0$, clearly

$$
\int_{0}^{\epsilon} \varphi(\mathrm{t}) \mathrm{dt}=\frac{\epsilon^{2}}{4}>0
$$

If $\mathrm{x}=\mathrm{y}$ for all $\mathrm{x}, \mathrm{y} \in \mathrm{E}$, then result holds trivially.
So suppose that $\mathrm{x} \neq \mathrm{y}$ for all $\mathrm{x}, \mathrm{y} \in \mathrm{E}$.
Since d is usual metric for all $\mathrm{x}, \mathrm{y} \in \mathrm{E}$, then on careful calculation, we get

$$
\text { L.H.S. }=\frac{|x-y|^{2}}{36}
$$

and

$$
\text { R.H.S. }=\frac{|x-y|^{2}}{3}
$$

Then clearly, L.H.S $\leq$ R.H.S for all $x, y \in E$ and, hence all conditions of Theorem 8 are verified.
Clearly $0 \in \mathrm{E}$ is the unique fixed point of A .

## References

[1] A. H. Ansari, Note on $\phi-\psi$ contractive type mappings and related fixed point, in: Proceedings of the 2nd Regional Conference on Mathematics And Applications, PNU, Iran, 2014, pp. 377-380.
[2] A. H. Ansari, S. Chandok and C. Ionescu, Fixed point theorems on bmetric spaces for weak contractions with auxiliary functions, J. Inequal. Appl., 2014 (2014), 1-17.
[3] A. Branciari, A fixed point theorem for mappings satisfying a general contractive condition of integral type, Int. J. Math. Math. Sci., 29 (2002), 531-536.
[4] B. K. Dass and S. Gupta, An extension of Banach contraction principle through rational expression, Indian J. Pure Appl. Math., 12 (1975), 14551458.
[5] B. E. Rhoades, Two fixed point theorems for mapping satisfying a general contractive condition of integral type, Int. J. Math. Math. Sci., 2003 (2003), 4007-4013.
[6] B. Samet and H. Yazidi, Fixed point theorems with respect to a contractive condition of integral type, Rend. Cric. Mat. Palermo., 60 (2011), 181-190.
[7] G. V. R. Babu and G. N. Alemayehu, Point of coincidence and common fixed points of a pair of generalized weakly contractive maps, Journal of Advanced Research in Pure Mathematics, 2 (2010), 89-106.
[8] N. Mani, Generalized $C_{\beta}^{\psi}$-rational contraction and fixed point theorem with application to second order differential equation, Mathematica Moravica, 22 (1) (2018), 43-54.
[9] N. Sharma, N. Mani, N. Gulati, Unique fixed point results for pairs of mappings on complete metric spaces, Ital. J. Pure Appl. Math., 42 (2) (2019), 798-808
[10] S. Banach, Sur les operations dans les ensembles abstraits et leur application aux equations integrals, Fundam. Math., 3 (1922), 133-181.
[11] S. M. A. Aleomraninejad and M. Shokouhnia, Some fixed point results of integral type and applications, Fixed Point Theory, 5 (2015), 101-109.
[12] S. Radenovic, A note on fixed point theory for cyclic phi -contractions, Fixed Point Theory Appl., 2015 (2015), 1-9.
[13] U. C. Gairola and A. S. Rawat, A fixed point theorem for two pair of maps satisfying a new contractive condition of integral type, Int. Math. Forum, 4 (2009), 177-183.
[14] V. Gupta and N. Mani, A common fixed point theorem for two weakly compatible mappings satisfying a new contractive condition of integral type, Mathematical Theory and Modeling, 1 (2011), 1-6.
[15] V. Gupta, N. Mani and A. K. Tripathi, A fixed point theorem satisfying a generalized weak contractive condition of integral type, Int. J. Math. Anal., Ruse, 6 (2012), 1883-1889.
[16] V. Gupta and N. Mani, Existence and uniqueness of fixed point for contractive mapping of integral type, Int. J. Comput. Sci. Math., 4 (2013), 72-83.
[17] V. Gupta, N. Mani, N. Sharma, Fixed point theorems for weak $(\psi, \beta)$ mappings satisfying generalized C -condition and its application to boundary value problem, Computational and Mathematical Methods, 1 (2019), e1041.
[18] Z. Q. Liu, X. Zou and S. M. Kang, Fixed point theorems of contractive mappings of integral type, Fixed Point Theory and Appl., 2013.

# Collatz conjecture revisited: an elementary generalization 

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#### Abstract

Collatz conjecture states that iterating the map that takes even natural number $n$ to $\frac{n}{2}$ and odd natural number $n$ to $3 n+1$, will eventually obtain 1 . In this paper a new generalization of the Collatz conjecture is analyzed and some interesting results are obtained. Since Collatz conjecture can be seen as a particular case of the generalization introduced in this articule, several more general conjectures are also presented.


## 1 Introduction

The Collatz conjecture remains today unsolved; as it has been for almost over 80 years. Although its statement is very simple and easy to understand, the nature of the problem makes extremelly to demonstrate or refuse. Articles such as [3] and [4] contain a huge amount of publications dealing with this problem and somehow trying to solve it.

Although the Collatz conjecture can be stated in several ways, in this paper we will use the following notation, i.e. modified Collatz function, that represents a slightly modification of the traditional formulation

$$
C(n)=\left\{\begin{array}{cl}
\frac{3 \cdot n+1}{2} & \text { if } n \equiv 1(\bmod 2) \\
\frac{n}{2} & \text { if } n \equiv 0((\bmod 2)
\end{array}\right.
$$

[^3]Key words and phrases: Collatz conjecture

By using the modified Collatz function, the Collatz conjecture can be stated in the following way:

Conjecture 1 (Collatz) For every integer number $\mathrm{n} \in \mathrm{N}$, there exists $k$ such that $\mathrm{C}^{(\mathrm{k})}(\mathrm{n})=1$.

In this sense, Terras [5] defined the total stopping time of an integer $n \in N$, here we denote it by $\sigma_{2}(n)$, as the smallest integer $k$ such that $C^{(k)}(n)=1$ or $\sigma_{2}(n)=\infty$ if no such $k$ exists. For example, if $n=7$, then by successively applying C , we obtain the following sequence $7 \rightarrow 11 \rightarrow 17 \rightarrow$ $26 \rightarrow 13 \rightarrow 20 \rightarrow 10 \rightarrow 5 \rightarrow 8 \rightarrow 4 \rightarrow 2 \rightarrow 1$, and then $C^{11}(7)=1$, so $\sigma_{2}(7)=11$.

In this paper, we consider an elementary generalization of the Collatz problem. We also generalize the concept of total stopping time and, related with this new approach, we obtain several results for some classes of integer numbers that generalize some well-known results on this topic. In the last section of the paper, some calculations and new conjectures are introduced. These conjectures try to illustrate the idea that traditional Collatz problem is just a special case and that it can be seen as part of a more general point of view, what we have called of Collatz numbers.

### 1.1 A more general $b$ - Collatz function, $C_{b}$

For any natural number $\mathrm{b} \geq 2$, we define the following $\mathrm{b}-$ Collatz function $\mathrm{C}_{\mathrm{b}}: \mathrm{N} \rightarrow \mathrm{N}$

$$
C_{b}(n)=\left\{\begin{array}{cl}
\frac{(b+1) \cdot n+(b-x)}{b} & \text { if } n \equiv x(\bmod b), 1 \leq x \leq b-1 \\
\frac{n}{b} & \text { if } n \equiv 0((\bmod b)
\end{array}\right.
$$

Clearly in the previous formula, if $\mathrm{b}=2$, we obtain the modified Collatz function stated in section [1]. In this sense, we defined the b-total stopping time of an integer $n \in N$, denoted $\sigma_{b}(n)$, as the smallest integer $k$ such that $C_{b}^{(k)}(\mathfrak{n})=1$ or $\sigma_{b}(\mathfrak{n})=\infty$ if no such $k$ exists. For example, if $b=5$ and $\mathrm{n}=7$, then by successively applying $\mathrm{C}_{5}$, we obtain the following sequence $7 \rightarrow 9 \rightarrow 11 \rightarrow 14 \rightarrow 17 \rightarrow 21 \rightarrow 26 \rightarrow 32 \rightarrow 39 \rightarrow 47 \rightarrow 57 \rightarrow 69 \rightarrow 83 \rightarrow$ $100 \rightarrow 20 \rightarrow 4 \rightarrow 5 \rightarrow 1$, so $\mathrm{C}_{5}^{17}(7)=1$ and finally $\sigma_{5}(7)=17$

Theorem 1 Let $\mathrm{b}, \mathrm{k}, \mathrm{r}, \mathrm{s} \in \mathrm{N}$, such that $\mathrm{b} \geq 2, \mathrm{k}, \mathrm{r} \geq 1, \mathrm{n}=\mathrm{b}^{\mathrm{k}} \cdot \mathrm{r}-\mathrm{s}>1$ and $1 \leq \mathrm{s} \leq \mathrm{b}-1$. Then

$$
C_{b}^{(k)}(n)=(b+1)^{k} \cdot r-s
$$

Proof. We proceed by induction on $t, 0 \leq t \leq k$. Case $t=0$.

$$
C_{b}^{(0)}\left(b^{k} \cdot r-s\right)=b^{k} \cdot r-s=\left((b+1)^{0} \cdot b^{k-0}\right) \cdot r-s
$$

So, the initial case holds. Now let us assume that

$$
C_{b}^{(t)}\left(b^{k} \cdot r-s\right)=\left((b+1)^{t} \cdot b^{k-t}\right) \cdot r-s
$$

for some $0 \leq t<k$. Since $t<k$, then $\left((b+1)^{t} \cdot b^{k-t}\right) \cdot r-s \equiv-s \equiv b-s(\bmod b)$, so

$$
\begin{aligned}
C_{b}^{(t+1)}\left(b^{k} \cdot r-s\right) & =C_{b}\left(C_{b}^{(t)}\left(b^{k} \cdot r-s\right)\right)=C_{b}\left(\left((b+1)^{t} \cdot b^{k-t}\right) \cdot r-s\right) \\
& =\frac{\left.(b+1) \cdot\left((b+1)^{t} \cdot b^{k-t}\right) \cdot r-s\right)+s}{b}= \\
& =\frac{(b+1)^{(t+1)} \cdot b^{k-t} \cdot r-b \cdot s}{b^{k}}= \\
& =(b+1)^{(t+1)} \cdot b^{k-(t+1)} \cdot r-s
\end{aligned}
$$

Thus,

$$
C_{b}^{(t+1)}\left(b^{k} \cdot r-s\right)=(b+1)^{(t+1)} \cdot b^{k-(t+1)} \cdot r-s
$$

and the result follows.

Corollary 1 If $b, k, r, s \in N$, such that $b \geq 2$, $k, r \geq 1, n=b^{k} \cdot r-s>1$ and $1 \leq s \leq b-1$, then

$$
\sigma_{\mathrm{b}}\left(\mathrm{~b}^{\mathrm{k}} \cdot \mathrm{r}-\mathrm{s}\right)=\sigma_{\mathrm{b}}\left((\mathrm{~b}+1)^{\mathrm{k}} \cdot \mathrm{r}-\mathrm{s}\right)+\mathrm{k}
$$

Corollary 2 If $\mathrm{b} \geq 2, \mathrm{k}, \mathrm{r} \geq 1, \mathrm{n}=\mathrm{b}^{\mathrm{k}}-1>1$, then

$$
\sigma_{b}\left(b^{k}-1\right)=\sigma_{b}\left((b+1)^{k}-1\right)+k
$$

Corollary 3 If $\mathrm{b}=2, \mathrm{k}, \mathrm{r} \geq 1, \mathrm{n}=2^{\mathrm{k}} \cdot \mathrm{r}-1$, then

$$
\sigma_{2}\left(2^{k} \cdot r-1\right)=\sigma_{2}\left(3^{k} \cdot r-1\right)+k
$$

Theorem 2 If $\mathrm{b}, \mathrm{k}, \mathrm{r}, \mathrm{t}, \mathrm{s} \in \mathrm{N}$, such that $\mathrm{b} \geq 2$, $\mathrm{k}>\mathrm{t} \geq 1, \mathrm{r} \geq 1$, and $1 \leq s \leq b-1$, then

$$
C_{b}^{(k)}\left(b^{k} \cdot r-b^{t} \cdot s\right)=(b+1)^{k-t} \cdot r-s
$$

Proof. Since $b^{k} \cdot r-b^{t} \cdot s$ is divisible by $b^{t}$, then $C_{b}^{(k)}\left(b^{k} \cdot r-b^{t} \cdot s\right)=C_{b}^{(k-t)}$ $\left(C_{b}^{t}\left(b^{k} \cdot r-b^{t} \cdot s\right)\right)=C_{b}^{(k-t)}\left(b^{k-t} \cdot r-s\right)=(b+1)^{k-t} \cdot r-s$, by Theorem 1 .

Corollary 4 If $\mathrm{b}, \mathrm{k}, \mathrm{r}, \mathrm{t}, \mathrm{s} \in \mathrm{N}$, such that $\mathrm{b} \geq 2, \mathrm{k}>\mathrm{t} \geq 1, \mathrm{r} \geq 1$, and $1 \leq s \leq b-1$, then

$$
\sigma_{\mathrm{b}}\left(\mathrm{~b}^{\mathrm{k}} \cdot \mathrm{r}-\mathrm{b}^{\mathrm{t}} \cdot \mathrm{~s}\right)=\sigma_{\mathrm{b}}\left((\mathrm{~b}+1)^{\mathrm{k}-\mathrm{t}} \cdot \mathrm{r}-\mathrm{s}\right)+\mathrm{k}
$$

Theorem 3 Let $\mathrm{b} \geq 2, \mathrm{k}>2$ and $\mathrm{r} \geq 1$, then
i) If $\mathrm{b}=2$ then $\sigma_{b}\left(\mathrm{~b}^{k} \cdot \mathrm{r}-(2 \mathrm{~b}-1)\right)=\sigma_{\mathrm{b}}\left((\mathrm{b}+1)^{\mathrm{k}-2} \cdot \mathrm{r}-1\right)+\mathrm{k}$.
ii) If $\mathrm{b}>2$ then $\sigma_{b}\left(\mathrm{~b}^{\mathrm{k}} \cdot \mathrm{r}-(2 \mathrm{~b}-1)\right)=\sigma_{\mathrm{b}}\left((\mathrm{b}+1)^{\mathrm{k}-1} \cdot \mathrm{r}-2\right)+\mathrm{k}$.

Proof. Let $n=b^{k} \cdot r-(2 b-1)$. Since $n \equiv 1(\bmod b)$, then

$$
\begin{aligned}
C_{b}^{(k)}(n) & =C_{b}^{(k-1)}\left(C_{b}(n)\right)=C_{b}^{(k-1)}\left(\frac{\left.(b+1) \cdot\left(b^{k} \cdot r-2 b+1\right)\right)+b-1}{b}\right) \\
& =C_{b}^{(k-1)}\left((b+1) \cdot b^{k-1} \cdot r-2 b\right)=C_{b}^{(k-2)}\left(C_{b}\left((b+1) \cdot b^{k-1} \cdot r-2 b\right)\right) \\
& =C_{b}^{(k-2)}\left((b+1) \cdot b^{k-2} \cdot r-2\right)
\end{aligned}
$$

If $b=2$, then $C_{b}^{(k)}(n)=C_{b}^{(k-2)}\left(3 \cdot 2^{k-2} \cdot r-2\right)=C_{b}^{(k-3)}\left(2^{k-3} \cdot 3 \cdot r-1\right)=3^{k-2} \cdot r-1$. If $b>2$, then by Theorem 1 ,

$$
C_{b}^{(k)}(n)=C_{b}^{(k-2)}\left((b+1) \cdot b^{k-2} \cdot r-2\right)=(b+1)^{k-1} \cdot r-2
$$

Corollary 5 [6] If $\mathrm{k}>2$ and $\mathrm{r} \geq 1$, then $\sigma_{2}\left(2^{\mathrm{k}} \cdot \mathrm{r}-3\right)=\sigma_{2}\left(3^{\mathrm{k}-2} \cdot \mathrm{r}-1\right)+\mathrm{k}$.
Theorem 4 Let $\mathrm{b} \geq 2, \mathrm{k}>2$ and $\mathrm{r}, \mathrm{t} \geq 1$, then
i) If $\mathrm{b}=2$, then $\sigma_{b}\left(\mathrm{~b}^{\mathrm{k}} \cdot \mathrm{r}-\mathrm{b}^{\mathrm{t}} \cdot(2 \mathrm{~b}-1)\right)=\sigma_{\mathrm{b}}\left((\mathrm{b}+1)^{(\mathrm{k}-\mathrm{t}-2)} \cdot \mathrm{r}-1\right)+\mathrm{k}$.
i) If $\mathrm{b}>2$, then $\sigma_{\mathrm{b}}\left(\mathrm{b}^{\mathrm{k}} \cdot \mathrm{r}-\mathrm{b}^{\mathrm{t}} \cdot(2 \mathrm{~b}-1)\right)=\sigma_{\mathrm{b}}\left((\mathrm{b}+1)^{(\mathrm{k}-\mathrm{t}-1)} \cdot \mathrm{r}-2\right)+\mathrm{k}$.

Proof. Let $n=b^{k} \cdot r-b^{t} \cdot(2 b-1)$. Since $n \equiv 0\left(\bmod b^{t}\right)$, then

$$
C_{b}^{(k)}(n)=C_{b}^{(k-t)}\left(\frac{n}{b^{t}}\right)=C_{b}^{k-t}\left(b^{k-t} \cdot r-(2 b-1)\right)
$$

By Theorem 3, if $\mathrm{b}=2$, then

$$
C_{b}^{(k)}(n)=C_{b}^{k-t}\left(b^{k-t} \cdot r-(2 b-1)\right)=(b+1)^{(k-t-2)} \cdot r-1
$$

and then

$$
\sigma_{b}\left(b^{k} \cdot r-b^{t} \cdot(2 b-1)\right)=\sigma_{b}\left((b+1)^{(k-t-2)} \cdot r-1\right)+k
$$

In othen hand, if $\mathrm{b}>2$, then

$$
C_{b}^{(k)}(n)=C_{b}^{k-t}\left(b^{k-t} \cdot r-(2 b-1)\right)=(b+1)^{(k-t-1)} \cdot r-2
$$

and then $\sigma_{b}\left(b^{k} \cdot r-b^{t} \cdot(2 b-1)\right)=\sigma_{b}\left((b+1)^{(k-t-1)} \cdot r-2\right)+k$.
Corollary 6 [6] If $\mathrm{k}>2$ and $\mathrm{r} \geq 1$, then $\sigma_{2}\left(2^{k} \cdot \mathrm{r}-6\right)=\sigma_{2}\left(3^{\mathrm{k}-3} \cdot \mathrm{r}-1\right)+\mathrm{k}$.
Theorem 5 Let $\mathrm{b} \geq 2$ and $\mathrm{k} \equiv 0(\bmod \mathrm{~b})$, then
i) $\sigma_{b}\left(b^{k}-1\right)=\sigma_{b}\left(b^{k}-1-1\right)+1$.
ii) $\sigma_{b}\left((b+1)^{k}-1\right)=\sigma_{b}\left((b+1)^{k-1}-1\right)$.

Proof. i) It enough to proof that

$$
C_{b}^{(k+1)}\left(b^{k-1}-1\right)=C_{b}^{(k+2)}\left(b^{k}-1\right)
$$

So, by Theorem $1, C_{b}^{(k+1)}\left(b^{k-1}-1\right)=C_{b}^{(2)}\left((b+1)^{k-1}-1\right)$. Since $(b+1)^{k-1}-1 \equiv$ $O(\bmod b)$, then

$$
C_{b}^{(2)}\left((b+1)^{k-1}-1\right)=C_{b}\left(\frac{(b+1)^{k-1}-1}{b}\right)
$$

Finally, since $k \equiv 0(\bmod b)$ and $\frac{(b+1)^{k-1}-1}{b} \equiv b-1(\bmod b)$, then

$$
C_{b}^{(2)}\left((b+1)^{k-1}-1\right)=C_{b}\left(\frac{(b+1)^{k-1}-1}{b}\right)=\frac{(b+1)^{k}-1}{b^{2}}
$$

In other hand, $C_{b}^{(k+2)}\left(b^{k}-1\right)=C_{b}^{(2)}\left((b+1)^{k}-1\right)$. So, since $k \equiv 0(\bmod b)$, then $(b+1)^{k}-1 \equiv 0\left(\bmod b^{2}\right)$. Therefore,

$$
C_{b}^{(k+2)}\left(b^{k}-1\right)=C_{b}^{(2)}\left((b+1)^{k}-1\right)=\frac{(b+1)^{k}-1}{b^{2}}
$$

and equality holds.
ii) From previous result i) and Corollary 4, we have

$$
\begin{aligned}
\sigma_{b}\left((b+1)^{k}-1\right) & =\sigma_{b}\left(b^{k}-1\right)-k=\sigma_{b}\left(b^{k-1}-1\right)+1-k \\
& =\sigma_{b}\left((b+1)^{k-1}-1\right)+(k-1)+1-k \\
& =\sigma_{b}\left((b+1)^{k-1}-1\right)
\end{aligned}
$$

Corollary 7 [6] If k is even and $\mathrm{k}>2$, then

$$
\begin{aligned}
& \sigma_{2}\left(2^{k} \cdot r-1\right)=\sigma_{2}\left(2^{k-1} \cdot r-1\right)+1, \text { and } \\
& \sigma_{2}\left(3^{k} \cdot r-1\right)=\sigma_{2}\left(3^{k-1} \cdot r-1\right) .
\end{aligned}
$$

## 3. Some empirical results on $C_{b}$ and open problems

The behaviour of $\mathrm{C}_{\mathrm{b}}$ has been intensively studied during last years when $\mathrm{b}=2$. But, what can we state for $C_{b}$ for other values of $b \geq 3$ ?. Let us see some examples for several values of $b$.

Case $b=3$. Calculating $C_{3}(n)$ for some values of $n$

| Iteration $/ \mathrm{n}=$ | 4 | 5 | 6 | 7 | 8 | 9 | 10 | 11 | 12 |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 6 | 7 | 2 | 10 | 11 | 3 | 14 | 15 | 4 |
| 2 | 2 | 10 | 3 | 14 | 15 | 1 | 19 | 5 | 6 |
| 3 | 3 | 14 | 1 | 19 | 5 | 2 | 26 | 7 | 2 |
| 4 | 1 | 19 | 2 | 26 | 7 | 3 | 35 | 10 | 3 |
| 5 | 2 | 26 | 3 | 35 | 10 | 1 | 47 | 14 | 1 |
| 5 | 3 | 35 | 1 | 47 | 14 | 2 | 63 | 19 | 2 |
| 6 | 1 | 47 | 2 | 63 | 19 | 3 | 21 | 26 | 3 |
| 7 | 2 | 63 | 3 | 21 | 26 | 1 | 7 | 35 | 1 |
| 8 | 3 | 21 | 1 | 7 | 35 | 2 | 10 | 47 | 2 |
| 9 | 1 | 7 | 2 | 10 | 47 | 3 | 14 | 63 | 3 |
| 10 | 2 | 10 | 3 | 14 | 63 | 1 | 19 | 21 | 1 |
| 11 | 3 | 14 | 1 | 19 | 21 | 2 | 26 | 7 | 2 |
| 12 | 1 | 19 | 2 | 26 | 7 | 3 | 35 | 10 | 3 |
| 13 | 2 | 26 | 3 | 35 | 10 | 1 | 47 | 14 | 1 |
| 14 | 3 | 35 | 1 | 47 | 14 | 2 | 63 | 19 | 2 |
| 15 | 1 | 47 | 2 | 63 | 19 | 3 | 21 | 26 | 3 |
| 16 | 2 | 63 | 3 | 21 | 26 | 1 | 7 | 35 | 1 |
| 17 | 3 | 21 | 1 | 7 | 35 | 2 | 10 | 47 | 2 |
| 18 | 1 | 7 | 2 | 10 | 47 | 3 | 14 | 63 | 3 |
| 19 |  |  |  |  |  |  |  |  |  |

At first sight, there is no homogeneous behaviour for different values of $n$. In this case, for $n \in\{4,6,9,12\}$ we found the cycle $2 \rightarrow 3 \rightarrow 1$ that repeats and contains the number 1 .

In other hand, for $\mathfrak{n} \in\{5,7,8,10,11\}$, the behaviour is completely different than the previous one. In this case, the cycle $7 \rightarrow 10 \rightarrow 14 \rightarrow 19 \rightarrow 26 \rightarrow$ $35 \rightarrow 47 \rightarrow 63 \rightarrow 21$ repeats and does not contain the number 1 . In fact, it seems that these are the only two cycles one can find when iterating $C_{3}$.

Case $b=5$. Calculating $C_{5}(n)$ for some values of $n$

| Iteration $/ \mathrm{n}=$ | 6 | 7 | 8 | 9 | 10 | 11 | 12 | 13 | 14 |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | ---: |
| 1 | 8 | 9 | 10 | 11 | 2 | 14 | 15 | 16 | 17 |
| 2 | 10 | 11 | 2 | 14 | 3 | 17 | 3 | 20 | 21 |
| 3 | 2 | 14 | 3 | 17 | 4 | 21 | 4 | 4 | 26 |
| 4 | 3 | 17 | 4 | 21 | 5 | 26 | 5 | 5 | 32 |
| 5 | 4 | 21 | 5 | 26 | 1 | 32 | 1 | 1 | 39 |
| 6 | 5 | 26 | 1 | 32 | 2 | 39 | 2 | 2 | 47 |
| 7 | 1 | 32 | 2 | 39 | 3 | 47 | 3 | 3 | 57 |
| $\mathbf{8}$ | 2 | 39 | 3 | 47 | 4 | 57 | 4 | 4 | 69 |
| 9 | 3 | 47 | 4 | 57 | 5 | 69 | 5 | 5 | 83 |
| 10 | 4 | 57 | 5 | 69 | 1 | 83 | 1 | 1 | 100 |
| 11 | 5 | 69 | 1 | 83 | 2 | 100 | 2 | 2 | 20 |
| 12 | 1 | 83 | 2 | 100 | 3 | 20 | 3 | 3 | 4 |
| 13 | 2 | 100 | 3 | 20 | 4 | 4 | 4 | 4 | 5 |
| 14 | 3 | 20 | 4 | 4 | 5 | 5 | 5 | 5 | 1 |
| 15 | 4 | 4 | 5 | 5 | 1 | 1 | 1 | 1 | 2 |
| 16 | 5 | 5 | 1 | 1 | 2 | 2 | 2 | 2 | 3 |
| 17 | 1 | 1 | 2 | 2 | 3 | 3 | 3 | 3 | 4 |
| 18 | 2 | 2 | 3 | 3 | 4 | 4 | 4 | 4 | 5 |
| 19 | 3 | 3 | 4 | 4 | 5 | 5 | 5 | 5 | 1 |
| 20 | 4 | 4 | 5 | 5 | 1 | 1 | 1 | 1 | 2 |
| 21 | 5 | 5 | 1 | 1 | 2 | 2 | 2 | 2 | 3 |

At first sight, there is an homogeneous behaviour for different values of $n$. Independently the value of $n$ we select, we found the cycle $2 \rightarrow 3 \rightarrow 4 \rightarrow 5 \rightarrow 1$ that repeats and contains the number 1.

Case $b=6$. Calculating $C_{6}(n)$ for some values of $n$

| Iteration/ $\mathrm{n}=$ | 7 | 8 | 9 | 10 | 11 | 12 | 13 | 14 | 15 |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 9 | 10 | 11 | 12 | 13 | 2 | 16 | 17 | 18 |
| 2 | 11 | 12 | 13 | 2 | 16 | 3 | 19 | 20 | 3 |
| 3 | 13 | 2 | 16 | 3 | 19 | 4 | 23 | 24 | 4 |
| 4 | 16 | 3 | 19 | 4 | 23 | 5 | 27 | 4 | 5 |
| 5 | 19 | 4 | 23 | 5 | 27 | 6 | 32 | 5 | 6 |
| 5 | 23 | 5 | 27 | 6 | 32 | 1 | 38 | 6 | 1 |
| 6 | 27 | 6 | 32 | 1 | 38 | 2 | 45 | 1 | 2 |
| 7 | 32 | 1 | 38 | 2 | 45 | 3 | 53 | 2 | 3 |
| 8 | 38 | 2 | 45 | 3 | 53 | 4 | 62 | 3 | 4 |
| 9 | 45 | 3 | 53 | 4 | 62 | 5 | 73 | 4 | 5 |
| 10 | 53 | 4 | 62 | 5 | 73 | 6 | 86 | 5 | 6 |
| 11 | 62 | 5 | 73 | 6 | 86 | 1 | 101 | 6 | 1 |
| 12 | 73 | 6 | 86 | 1 | 101 | 2 | 118 | 1 | 2 |
| 13 | 86 | 1 | 101 | 2 | 118 | 3 | 138 | 2 | 3 |
| 14 | 101 | 2 | 118 | 3 | 138 | 4 | 23 | 3 | 4 |
| 15 | 118 | 3 | 138 | 4 | 23 | 5 | 27 | 4 | 5 |
| 16 | 138 | 4 | 23 | 5 | 27 | 6 | 32 | 5 | 6 |
| 17 | 23 | 5 | 27 | 6 | 32 | 1 | 38 | 6 | 1 |
| 18 |  |  |  |  |  |  |  |  |  |

By inspectioning this examples, it can be seen that there are two groups of numbers. First, numbers such as $n \in\{7,9,11,13\}$ whose iterations contain the cycle $23 \rightarrow 27 \rightarrow 32 \rightarrow 38 \rightarrow 45 \rightarrow 53 \rightarrow 62 \rightarrow 73 \rightarrow 86 \rightarrow 101 \rightarrow 118 \rightarrow 138$. Secondly, numbers such as $n \in\{8,10,12,14,15\}$ whose iterations contain the cycle $2 \rightarrow 3 \rightarrow 4 \rightarrow 5 \rightarrow 6 \rightarrow 1$ that repeats and contains the number 1 .

Case $b=7$. Calculating $C_{7}(n)$ for some values of $n$

| Iteration/ $\mathrm{n}=$ | 8 | 9 | 10 | 11 | 12 | 13 | 14 | 15 | 16 |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 10 | 11 | 12 | 13 | 14 | 15 | 2 | 18 | 19 |
| 2 | 12 | 13 | 14 | 15 | 2 | 18 | 3 | 21 | 22 |
| 3 | 14 | 15 | 2 | 18 | 3 | 21 | 4 | 3 | 26 |
| 4 | 2 | 18 | 3 | 21 | 4 | 3 | 5 | 4 | 30 |
| 5 | 3 | 21 | 4 | 3 | 5 | 4 | 6 | 5 | 35 |
| 6 | 4 | 3 | 5 | 4 | 6 | 5 | 7 | 6 | 5 |
| 7 | 5 | 4 | 6 | 5 | 7 | 6 | 1 | 7 | 6 |
| 8 | 6 | 5 | 7 | 6 | 1 | 7 | 2 | 1 | 7 |
| 9 | 7 | 6 | 1 | 7 | 2 | 1 | 3 | 2 | 1 |
| 10 | 1 | 7 | 2 | 1 | 3 | 2 | 4 | 3 | 2 |
| 11 | 2 | 1 | 3 | 2 | 4 | 3 | 5 | 4 | 3 |
| 12 | 3 | 2 | 4 | 3 | 5 | 4 | 6 | 5 | 4 |
| 13 | 4 | 3 | 5 | 4 | 6 | 5 | 7 | 6 | 5 |
| 14 | 5 | 4 | 6 | 5 | 7 | 6 | 1 | 7 | 6 |
| 15 | 6 | 5 | 7 | 6 | 1 | 7 | 2 | 1 | 7 |
| 16 | 7 | 6 | 1 | 7 | 2 | 1 | 3 | 2 | 1 |
| 17 | 1 | 7 | 2 | 1 | 3 | 2 | 4 | 3 | 2 |
| 18 | 2 | 1 | 3 | 2 | 4 | 3 | 5 | 4 | 3 |

By inspectioning this examples, it can be seen that in all cases, iterations contain the cycle $2 \rightarrow 3 \rightarrow 4 \rightarrow 5 \rightarrow 6 \rightarrow 7 \rightarrow 1$ that repeats and contains the number 1 .

Case $b=9$. Calculating $C_{9}(n)$ for some values of $n$

| Iteration/ $\mathrm{n}=$ | 10 | 11 | 12 | 13 | 14 | 15 | 16 | 17 | 31 | 35 |
| :--- | :--- | :--- | :--- | :--- | :--- | :---: | :---: | :---: | :---: | :---: |
| 1 | 12 | 13 | 14 | 15 | 16 | 17 | 18 | 19 | 35 | 39 |
| 2 | 14 | 15 | 16 | 17 | 18 | 19 | 2 | 22 | 39 | 44 |
| 3 | 16 | 17 | 18 | 19 | 2 | 22 | 3 | 25 | 44 | 49 |
| 4 | 18 | 19 | 2 | 22 | 3 | 25 | 4 | 28 | 49 | 55 |
| 5 | 2 | 22 | 3 | 25 | 4 | 28 | 5 | 32 | 55 | 62 |
| 6 | 3 | 25 | 4 | 28 | 5 | 32 | 6 | 36 | 62 | 69 |
| 7 | 4 | 28 | 5 | 32 | 6 | 36 | 7 | 4 | 69 | 77 |
| 8 | 5 | 32 | 6 | 36 | 7 | 4 | 8 | 5 | 77 | 86 |
| 9 | 6 | 36 | 7 | 4 | 8 | 5 | 9 | 6 | 86 | 96 |
| 10 | 7 | 4 | 8 | 5 | 9 | 6 | 1 | 7 | 96 | 107 |
| 11 | 8 | 5 | 9 | 6 | 1 | 7 | 2 | 8 | 107 | 119 |
| 12 | 9 | 6 | 1 | 7 | 2 | 8 | 3 | 9 | 119 | 133 |
| 13 | 1 | 7 | 2 | 8 | 3 | 9 | 4 | 1 | 133 | 148 |
| 14 | 2 | 8 | 3 | 9 | 4 | 1 | 5 | 2 | 148 | 165 |
| 15 | 3 | 9 | 4 | 1 | 5 | 2 | 6 | 3 | 165 | 184 |
| 16 | 4 | 1 | 5 | 2 | 6 | 3 | 7 | 4 | 184 | 205 |
| 17 | 5 | 2 | 6 | 3 | 7 | 4 | 8 | 5 | 205 | 228 |
| 18 | 6 | 3 | 7 | 4 | 8 | 5 | 9 | 6 | 228 | 254 |
| 19 | 7 | 4 | 8 | 5 | 9 | 6 | 1 | 7 | 254 | 283 |
| 20 | 8 | 5 | 9 | 6 | 1 | 7 | 2 | 8 | 283 | 315 |
| 21 | 9 | 6 | 1 | 7 | 2 | 8 | 3 | 9 | 315 | 35 |
| 22 | 1 | 7 | 2 | 8 | 3 | 9 | 4 | 1 | 35 | 39 |
| 23 | 2 | 8 | 3 | 9 | 4 | 1 | 5 | 2 | 39 | 44 |
| 24 | 3 | 9 | 4 | 1 | 5 | 2 | 6 | 3 | 44 | 49 |
| 25 | 4 | 1 | 5 | 2 | 6 | 3 | 7 | 4 | 49 | 55 |

By inspectioning this examples, it can be seen that for $n \in\{10,11,12,13,14$, $15,16,17\}$, iterations contain the cycle $2 \rightarrow 3 \rightarrow 4 \rightarrow 5 \rightarrow 6 \rightarrow 7 \rightarrow 8 \rightarrow 9 \rightarrow 1$ that repeats and contains the number 1 . In other hand, for $n \in\{31,35\}$, for example, interations contain the cycle $35 \rightarrow 39 \rightarrow 44 \rightarrow 49 \rightarrow 55 \rightarrow 62 \rightarrow$ $69 \rightarrow 77 \rightarrow 86 \rightarrow 96 \rightarrow 107 \rightarrow 119 \rightarrow 133 \rightarrow 148 \rightarrow 165 \rightarrow 184 \rightarrow 205 \rightarrow$ $228 \rightarrow 254 \rightarrow 283 \rightarrow 315$.

Case $b=10$. Calculating $C_{10}(n)$ for some values of $n$. We compute the fisrt 30 interations for esch value of $n$,

| Iteration $/ \mathrm{n}=$ | 11 | 12 | 13 | 14 | 15 | 16 | 17 | 18 | 34 | 38 |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 13 | 14 | 15 | 16 | 17 | 18 | 19 | 20 | 38 | 42 |
| 2 | 15 | 16 | 17 | 18 | 19 | 20 | 21 | 2 | 42 | 47 |
| 3 | 17 | 18 | 19 | 20 | 21 | 2 | 24 | 3 | 47 | 52 |
| 4 | 19 | 20 | 21 | 2 | 24 | 3 | 27 | 4 | 52 | 58 |
| 5 | 21 | 2 | 24 | 3 | 27 | 4 | 30 | 5 | 58 | 64 |
| 6 | 24 | 3 | 27 | 4 | 30 | 5 | 3 | 6 | 64 | 71 |
| 7 | 27 | 4 | 30 | 5 | 3 | 6 | 4 | 7 | 71 | 79 |
| 8 | 30 | 5 | 3 | 6 | 4 | 7 | 5 | 8 | 79 | 87 |
| 9 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 | 87 | 96 |
| 10 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 | 96 | 106 |
| 11 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 1 | 106 | 117 |
| 12 | 6 | 9 | 7 | 10 | 8 | 1 | 9 | 2 | 117 | 129 |
| 13 | 7 | 10 | 8 | 1 | 9 | 2 | 10 | 3 | 129 | 142 |
| 14 | 8 | 1 | 9 | 2 | 10 | 3 | 1 | 4 | 142 | 157 |
| 15 | 9 | 2 | 10 | 3 | 1 | 4 | 2 | 5 | 157 | 173 |
| 16 | 10 | 3 | 1 | 4 | 2 | 5 | 3 | 6 | 173 | 191 |
| 17 | 1 | 4 | 2 | 5 | 3 | 6 | 4 | 7 | 191 | 211 |
| 18 | 2 | 5 | 3 | 6 | 4 | 7 | 5 | 8 | 211 | 233 |
| 19 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 | 233 | 257 |
| 20 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 | 257 | 283 |
| 21 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 1 | 283 | 312 |
| 22 | 6 | 9 | 7 | 10 | 8 | 1 | 9 | 2 | 312 | 344 |
| 23 | 7 | 10 | 8 | 1 | 9 | 2 | 10 | 3 | 344 | 379 |
| 24 | 8 | 1 | 9 | 2 | 10 | 3 | 1 | 4 | 379 | 417 |
| 25 | 9 | 2 | 10 | 3 | 1 | 4 | 2 | 5 | 417 | 459 |
| 26 | 10 | 3 | 1 | 4 | 2 | 5 | 3 | 6 | 459 | 505 |
| 27 | 1 | 4 | 2 | 5 | 3 | 6 | 4 | 7 | 505 | 556 |
| 28 | 2 | 5 | 3 | 6 | 4 | 7 | 5 | 8 | 556 | 612 |
| 29 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 | 612 | 674 |
| 30 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 | 674 | 742 |

Then, we compute values for the rest of interations

| Iteration $/ \mathrm{n}=$ | 11 | 12 | 13 | 14 | 15 | 16 | 17 | 18 | 34 | 38 |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | ---: |
| 31 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 1 | 742 | 817 |
| 32 | 6 | 9 | 7 | 10 | 8 | 1 | 9 | 2 | 817 | 899 |
| 33 | 7 | 10 | 8 | 1 | 9 | 2 | 10 | 3 | 899 | 989 |
| 34 | 8 | 1 | 9 | 2 | 10 | 3 | 1 | 4 | 989 | 1088 |
| 35 | 9 | 2 | 10 | 3 | 1 | 4 | 2 | 5 | 1088 | 1197 |
| 36 | 10 | 3 | 1 | 4 | 2 | 5 | 3 | 6 | 1197 | 1317 |
| 37 | 1 | 4 | 2 | 5 | 3 | 6 | 4 | 7 | 1317 | 1449 |
| 38 | 2 | 5 | 3 | 6 | 4 | 7 | 5 | 8 | 1449 | 1594 |
| 39 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 | 1594 | 1754 |
| 40 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 | 1754 | 1930 |
| 41 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 1 | 1930 | 193 |
| 42 | 6 | 9 | 7 | 10 | 8 | 1 | 9 | 2 | 193 | 213 |
| 43 | 7 | 10 | 8 | 1 | 9 | 2 | 10 | 3 | 213 | 235 |
| 44 | 8 | 1 | 9 | 2 | 10 | 3 | 1 | 4 | 235 | 259 |
| 45 | 9 | 2 | 10 | 3 | 1 | 4 | 2 | 5 | 259 | 285 |
| 46 | 10 | 3 | 1 | 4 | 2 | 5 | 3 | 6 | 285 | 314 |
| 47 | 1 | 4 | 2 | 5 | 3 | 6 | 4 | 7 | 314 | 346 |
| 48 | 2 | 5 | 3 | 6 | 4 | 7 | 5 | 8 | 346 | 381 |
| 49 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 | 381 | 420 |
| 50 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 | 420 | 42 |
| 51 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 1 | 42 | 47 |

In this case, for $n \in\{11,12,13,14,15,16,17,18\}$ there is a common cycle $2 \rightarrow 3 \rightarrow 4 \rightarrow 5 \rightarrow 6 \rightarrow 7 \rightarrow 8 \rightarrow 9 \rightarrow 10 \rightarrow 1$ that repeats and contains the number 1 . In other hand, for numbers such as $n \in\{34,38\}$, one can find the large cycle $42 \rightarrow 47 \rightarrow 52 \rightarrow 58 \rightarrow 64 \rightarrow 71 \rightarrow 79 \rightarrow 87 \rightarrow 96 \rightarrow$ $106 \rightarrow 117 \rightarrow 129 \rightarrow 142 \rightarrow 157 \rightarrow 173 \rightarrow 191 \rightarrow 211 \rightarrow 233 \rightarrow 257 \rightarrow 283 \rightarrow$ $312 \rightarrow 344 \rightarrow 379 \rightarrow 417 \rightarrow 459 \rightarrow 505 \rightarrow 556 \rightarrow 612 \rightarrow 674 \rightarrow 742 \rightarrow 817 \rightarrow$ $899 \rightarrow 1088 \rightarrow 1197 \rightarrow 1317 \rightarrow 1449 \rightarrow 1594 \rightarrow 1754 \rightarrow 1930 \rightarrow 193 \rightarrow 213 \rightarrow$ $235 \rightarrow 259 \rightarrow 285 \rightarrow 314 \rightarrow 346 \rightarrow 381 \rightarrow 420$ that repeats and does not contain the number 1 .

Case $b=11$. Calculating $C_{11}(n)$ for some values of $n$

| Iteration/n $=$ | 12 | 13 | 14 | 15 | 16 | 17 | 18 | 19 | 20 | 21 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 14 | 15 | 16 | 17 | 18 | 19 | 20 | 21 | 22 | 23 |
| 2 | 16 | 17 | 18 | 19 | 20 | 21 | 22 | 23 | 2 | 26 |
| 3 | 18 | 19 | 20 | 21 | 22 | 23 | 2 | 26 | 3 | 29 |
| 4 | 20 | 21 | 22 | 23 | 2 | 26 | 3 | 29 | 4 | 32 |
| 5 | 22 | 23 | 2 | 26 | 3 | 29 | 4 | 32 | 5 | 35 |
| 6 | 2 | 26 | 3 | 29 | 4 | 32 | 5 | 35 | 6 | 39 |
| 7 | 3 | 29 | 4 | 32 | 5 | 35 | 6 | 39 | 7 | 43 |
| 8 | 4 | 32 | 5 | 35 | 6 | 39 | 7 | 43 | 8 | 47 |
| 9 | 5 | 35 | 6 | 39 | 7 | 43 | 8 | 47 | 9 | 52 |
| 10 | 6 | 39 | 7 | 43 | 8 | 47 | 9 | 52 | 10 | 57 |
| 11 | 7 | 43 | 8 | 47 | 9 | 52 | 10 | 57 | 11 | 63 |
| 12 | 8 | 47 | 9 | 52 | 10 | 57 | 11 | 63 | 1 | 69 |
| 13 | 9 | 52 | 10 | 57 | 11 | 63 | 1 | 69 | 2 | 76 |
| 14 | 10 | 57 | 11 | 63 | 1 | 69 | 2 | 76 | 3 | 83 |
| 15 | 11 | 63 | 1 | 69 | 2 | 76 | 3 | 83 | 4 | 91 |
| 16 | 1 | 69 | 2 | 76 | 3 | 83 | 4 | 91 | 5 | 100 |
| 17 | 2 | 76 | 3 | 83 | 4 | 91 | 5 | 100 | 6 | 110 |
| 18 | 3 | 83 | 4 | 91 | 5 | 100 | 6 | 110 | 7 | 10 |
| 19 | 4 | 91 | 5 | 100 | 6 | 110 | 7 | 10 | 8 | 11 |
| 20 | 5 | 100 | 6 | 110 | 7 | 10 | 8 | 11 | 9 | 1 |
| 21 | 6 | 110 | 7 | 10 | 8 | 11 | 9 | 1 | 10 | 2 |
| 22 | 7 | 10 | 8 | 11 | 9 | 1 | 10 | 2 | 11 | 3 |
| 23 | 8 | 11 | 9 | 1 | 10 | 2 | 11 | 3 | 1 | 4 |
| 24 | 9 | 1 | 10 | 2 | 11 | 3 | 1 | 4 | 2 | 5 |
| 25 | 10 | 2 | 11 | 3 | 1 | 4 | 2 | 5 | 3 | 6 |
| 26 | 11 | 3 | 1 | 4 | 2 | 5 | 3 | 6 | 4 | 7 |
| 27 | 1 | 4 | 2 | 5 | 3 | 6 | 4 | 7 | 5 | 8 |
| 28 | 2 | 5 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 |
| 29 | 3 | 6 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 |
| 30 | 4 | 7 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 11 |
| 31 | 5 | 8 | 6 | 9 | 7 | 10 | 8 | 11 | 9 | 1 |
| 32 | 6 | 9 | 7 | 10 | 8 | 11 | 9 | 1 | 10 | 2 |
| 33 | 7 | 10 | 8 | 11 | 9 | 1 | 10 | 2 | 11 | 3 |
| 34 | 8 | 11 | 9 | 1 | 10 | 2 | 11 | 3 | 1 | 4 |
| 35 | 9 | 1 | 10 | 2 | 11 | 3 | 1 | 4 | 2 | 5 |

By inspectioning this examples, it can be seen that in all cases, iterations contain the cycle $2 \rightarrow 3 \rightarrow 4 \rightarrow 5 \rightarrow 6 \rightarrow 7 \rightarrow 8 \rightarrow 9 \rightarrow 10 \rightarrow 11 \rightarrow 1$ that repeats and contains the number 1.

From previous empirical results, following definitions and conjectures can be stated

Definition 1 For any $b \geq 2$, set $\{2,3, \ldots, b, 1\}$ is called $a b$-trivial cycle.
Definition 2 Let $\mathrm{b} \geq 2$ an integer number. Then, for any $\mathrm{n} \in \mathrm{N}$, we define

$$
\operatorname{Iter}_{\mathrm{b}}(\mathrm{n})=\left\{\mathrm{C}_{\mathrm{b}}^{(\mathrm{k})}(\mathrm{n}) \mid \mathrm{k} \geq 1\right\}
$$

Definition 3 An integer number $\mathrm{b} \geq 2$ is called Collatz number if for every $\mathrm{n} \in \mathrm{N}$, $\operatorname{Iter}_{\mathrm{b}}(\mathrm{n})$ contains the b -trivial cycle.

Lemma 1 Let $\mathrm{b} \geq 2$ an integer number. Then, there exist numbers $\mathrm{n} \in \mathrm{N}$, such that
i) $1 \in \operatorname{Iter}_{\mathrm{b}}(\mathrm{n})$, and
ii) $\{2,3, \ldots, b, 1\} \subset \operatorname{Iter}_{b}(n)$.

Proof. Let $n=2 \cdot b$, then it is easy to verify that $C_{b}(n)=2, C_{b}^{(2)}(n)=3, \ldots$, $C_{b}^{b-1}(n)=b, C_{b}^{(b)}(n)=1$ and $C_{b}^{(b+1)}(n)=2$. Thus, both previous statements can be easily checked.

If we analyze the list of positive integers numbers lower than 20, then we can see two completely different behaviour when iterating $C_{b}$ function. In one hand, we can find values for $b$, such as $2,5,7,8, \ldots$, in which interations of $C_{b}(n)$ on any integer number $n$, it seems, always end in 1 , more concretely, interations contain the $b$-trivial cycle. And in other hand, values of $b$, such as $3,4,6,9,10, \ldots$, in which interations of $C_{b}(n)$ on any integer number $n$ either end in 1 or in another non trivial cycle. Below you can find a table for different values of $b$, where one can find the lowest value of $n$, for which $\operatorname{Iter}_{b}(n)$ does not contain 1, but however, it contains a non trivial cycle.

| b | n |
| :--- | ---: |
| 3 | 5 |
| 4 | 11 |
| 6 | 7 |
| 9 | 31 |
| 10 | 34 |
| 11 | 588 |
| 12 | 767 |
| 15 | 49 |
| 16 | 35 |
| 17 | 19 |

Conjecture 2 Let $\mathrm{b} \geq 2$ an integer number. Then, one can find only following two possibilities

Case i) For all $\mathrm{n} \in \mathrm{N}$, $\operatorname{Iter}_{\mathrm{b}}(\mathrm{n})$ contains the b -trivial cycle, and then b is a Collatz number, or

Case ii) There are $\mathrm{n} \in \mathrm{N}$, such that $\mathrm{Iter}_{\mathrm{b}}(\mathrm{n})$ contains a common non-trivial cycle. There are also other values of $\mathrm{n} \in \mathrm{N}$, for which $\operatorname{Iter}_{\mathrm{b}}(\mathrm{n})$ contains the b-trivial cycle.

Conjecture 3 Numbers 2, 5, 7, 8, 13, 14, 18 and 19 are Collatz numbers.
Conjecture 4 There are infinite Collatz numbers.

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## References

[1] L. Collatz, On the motivation and origin of the (3n+1)-problem, J. of Qufu Normal University, Natural Science Edition, 12 (1986), 9-11.
[2] J. Lagarias, The $3 x+1$ problem and its generalizations, American Mathematical Monthly, 92:3-23, 1985.
[3] J. Lagarias, The $3 x+1$ problem: an annotated bibliography I (1963-1999), http://arxiv.org/abs/ math/0309224v13, 2011.
[4] J. Lagarias, The $3 x+1$ problem: an annotated bibliography II (2000-2009), http://arxiv.org/ abs/math/0608208v6, 2012.
[5] R. Terras, A stopping time problem on the positive integers, Acta Arth., 30 (3) (1976), 241-252.
[6] P. Wiltrout, E. Landquist, The Collatz conjecture and integers of the form $2^{k} b-m$ and $3^{k} b-1$, Furman University,. Electronic Journal of Undergraduate Mathematics, 17 (2013), 1-5.

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# Nonparametric estimation of trend function for stochastic differential equations driven by a bifractional Brownian motion 

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Abstract. The main objective of this paper is to investigate the problem of estimating the trend function $S_{t}=S\left(x_{t}\right)$ for process satisfying stochastic differential equations of the type

$$
\mathrm{d} X_{\mathrm{t}}=\mathrm{S}\left(\mathrm{X}_{\mathrm{t}}\right) \mathrm{dt}+\varepsilon \mathrm{dB}_{\mathrm{t}}^{\mathrm{H}, \mathrm{~K}}, \mathrm{X}_{0}=\mathrm{x}_{0}, 0 \leq \mathrm{t} \leq \mathrm{T}
$$

where $\left\{B_{t}^{H, K}, t \geq 0\right\}$ is a bifractional Brownian motion with known parameters $H \in(0,1), K \in(0,1]$ and $H K \in(1 / 2,1)$. We estimate the unknown function $S\left(x_{t}\right)$ by a kernel estimator $\widehat{S}_{t}$ and obtain the asymptotic properties as $\varepsilon \longrightarrow 0$. Finally, a numerical example is provided.

[^4]
## 1 Introduction

Fractional Brownian motion (fBm) is the most well-known and employed process with a long dependency-property for many real world applications including telecommunication, turbulence, finance, and so on. This process was introduced by Kolmogorov [5], then studied by many researchers including Mandelbrot and Van Ness [9] and Norros et al. [12].

The bifractional Brownian motion (bfBm) was introduced in Houdré and Villa [3], and further studied by Russo and Tudor [14] and Tudor and Xiao [16].

Nonparametric estimation of trend function for stochastic differential equations (SDEs) has caught the attention of different researchers. It was first investigated by Kutoyants [7] for the stochastic differential equation driven by a standard Brownian motion. After that, the problem was generalized by Mishra and Rao [10] for the stochastic differential equation driven by a fractional Brownian motion. Then, Mishra and Rao [11] presented nonparametric estimation of linear multiplier for fractional diffusion processes. Later, nonparametric inference for fractional diffusion were dealt by Saussereau [15]. Very recently, Prakasa Rao [13] investigated nonparametric estimation of trend function for SDEs driven by mixed fractional Brownian motion.

In this paper, we use the method developed by Kutoyants [7] to construct an estimate of the trend function $S_{t}$ in a model described by stochastic differential equations driven by a bifractional Brownian motion. For this, let $\left\{X_{t}, 0 \leq t \leq T\right\}$ be the process governed by the following equation:

$$
\mathrm{d} \mathrm{X}_{\mathrm{t}}=\mathrm{S}\left(\mathrm{X}_{\mathrm{t}}\right) \mathrm{dt}+\varepsilon \mathrm{dB}_{\mathrm{t}}^{\mathrm{H}, \mathrm{~K}}, \mathrm{X}_{0}=x_{0}, 0 \leq \mathrm{t} \leq \mathrm{T},
$$

where $\varepsilon>0$ and $\mathrm{B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{K}}$ is a bifractional Brownian motion of parameters $\mathrm{H} \in$ $(0,1), K \in(0,1]$, and $S($.$) is an unknown function. In Kutoyants [7]$, the trend coefficient in a diffusion process was estimated from the process $\left\{X_{t}, 0 \leq t \leq T\right\}$. In this investigation, we use a similar approach and consider the estimate $\hat{S}_{t}$ of $S_{t}$ as follows:

$$
\hat{S}_{t}=\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d X_{\tau},
$$

where $G$ is a bounded kernel with finite support with $\phi_{\varepsilon} \longrightarrow 0$ as $\varepsilon \longrightarrow 0$. Under some hypotheses, we firstly prove the mean square consistency of the estimator. Then, we give a bound on the rate of convergence and prove the asymptotic normality of the estimator $\hat{S}_{t}$.

To the best of our knowledge, the problem of nonparametric estimation of
trend function for stochastic differential equations driven by a bfBm has not been considered in the literature.

The rest of the paper is structured as follows. In Section 2, the basic properties of bifractional Brownian motion are stated. Section 3 is devoted to the preliminaries. Then, in Section 4, we give the main results; under some hypotheses, we establish the uniform consistency (Theorem 1), the rate of convergence (Theorem 2) as well as the asymptotic normality (Theorem 3) of the estimator. Further, in Section 5, a simulation example is carried out to illuminate our theoretical study. Section 6 is devoted to the technical proofs. Finally, we conclude the paper in Section 7.

## 2 Bifractional Brownian motion

Let $\left(\Omega, \mathcal{F},\left\{\mathcal{F}_{\mathrm{t}}\right\}_{\mathrm{t}} \geq 0, \mathbb{P}\right)$ be a stochastic basis satisfying the habitual hypotheses, i.e., a filtered probability space with a right continuous filtration $\left\{\mathcal{F}_{\mathrm{t}}\right\}_{\mathrm{t} \geq 0}$ and $\mathcal{F}_{0}$ contains every $\mathbb{P}$-null set.

Let $\left\{\mathrm{B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{K}}, \mathrm{t} \geq 0\right\}$ be a normalized bifractional Brownian motion with parameters $\mathrm{H} \in(0,1)$ and $\mathrm{K} \in(0,1]$, that is, a Gaussian process with continuous sample paths with $\mathrm{B}_{0}^{\mathrm{H}, \mathrm{K}}=0$ and the covariance:

$$
R_{H, K}(t, s)=\mathbb{E}\left(B_{t}^{H, K} B_{s}^{H, K}\right)=\frac{1}{2^{K}}\left[\left(t^{2 H}+s^{2 H}\right)^{K}+|s-t|^{2 H K}\right], \quad t \geq 0, s \geq 0 .
$$

When $\mathrm{K}=1$, we retrieve the fractional Brownian motion while the case $\mathrm{K}=1$ and $\mathrm{H}=1 / 2$ corresponds to the standard Brownian motion.

The bfBm is an extension of the fBm which preserves many properties of the fBm , but not the stationarity of the increments. Russo and Tudor [14] showed that the $\mathrm{bfBm} \mathrm{B}^{\mathrm{H}, \mathrm{K}}$ behaves as a fBm of Hurst parameter HK.

According to Houdré and Villa [3] and Tudor and Xiao [16], the bfBm has the following properties:

1. $\mathbb{E}\left(\mathrm{B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{H}}\right)=0$ and $\operatorname{Var}\left(\mathrm{B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{K}}\right)=\mathrm{t}^{2 \mathrm{HK}}$.
2. $\mathrm{B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{K}}$ is said to be self-similar with index $\mathrm{HK} \in(0,1)$, that is, for every constant $a>0$,

$$
\begin{equation*}
\left\{\mathrm{B}_{\mathrm{at}}^{\mathrm{H}, \mathrm{~K}}, \mathrm{t} \geq 0\right\} \stackrel{\Delta}{=}\left\{\mathrm{a}^{\mathrm{HK}} \mathrm{~B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{~K}}, \mathrm{t} \geq 0\right\} \text {, for each } \mathrm{a}>0 \tag{1}
\end{equation*}
$$

in the sense that the processes, on both sides of the equality sign, have the same finite dimensional distributions.
3. The process $B_{t}^{H, K}$ is not Markov and it is not a semi-martingale if $H K \neq$ $1 / 2$.
4. The trajectories of the process $\mathrm{B}^{\mathrm{H}, \mathrm{K}}$ are Hölder continuous of order $\delta$ for any $\delta<\mathrm{HK}$ and they are nowhere differentiable.
5. The $\operatorname{bfBm} B^{H, K}$ is a quasi-helix in the sense of Kahane [4], for any $t, s \geq 0$ we have

$$
2^{-K}(t-s)^{2 H K} \leq \mathbb{E}\left[B_{t}^{H, K}-B_{s}^{H, K}\right]^{2} \leq 2^{1-K}(t-s)^{2 H K}
$$

The $\operatorname{bfBm} B^{H, K}$ can be extended for $K \in(1,2)$ with $H \in(0,1)$ and $H K \in(0,1)$ (see Bardina and Es-Sebaiy [1] and Lifshits and Volkova [8]).

The stochastic calculus with respect to the bifractional Brownian motion has been recently developed by Kruk et al. [6]. More works on bifractional Brownian motion can be found in Tudor and Xiao [16], Es-sabaiy and Tudor [2], Yan et al. [17] and the references therein.

Fix a time interval $[0, \mathrm{~T}]$, we denote by $\mathcal{E}$ the set of step function on $[0, T]$. Let $\mathcal{H}_{\mathrm{B}}{ }^{\boldsymbol{\prime}, \mathrm{K}}$ be the canonical Hilbert space associated to the bfBm defined as the closure of $\mathcal{E}$ with respect to the scalar product

$$
\left\langle 1_{[0, t]}, 1_{[0, s]}\right\rangle_{\mathcal{H}_{B} H, K}=R_{H, K}(t, s)=\int_{0}^{T} \int_{0}^{T} 1_{[0, t]}(u) 1_{[0, s]}(v) \frac{\partial^{2} R_{H, K}(u, v)}{\partial u \partial v} d u d v
$$

where $R_{H, K}(t, s)$ is the covariance of $B_{t}^{H, K}$ and $B_{s}^{H, K}$. The application $\varphi \in \mathcal{E} \longrightarrow$ $B^{\mathrm{H}, \mathrm{K}}(\varphi)$ is an isometry from $\mathcal{E}$ to the Gaussian space generated by $\mathrm{B}^{\mathrm{H}, \mathrm{K}}$ and it can be extended to $\mathcal{H}_{\mathrm{B}}{ }^{\mu, \kappa}$. In this study, as $\operatorname{HK} \in(1 / 2,1)$ we will employ the subspace $\left|\mathcal{H}_{\mathrm{BH}, \mathrm{K}}\right|$ of $\mathcal{H}_{\mathrm{B}, \mathrm{K}}$ which is defined as the set of measurable function $\varphi$ on $[0, T]$ satisfying

$$
\begin{equation*}
\|\varphi\|_{\left|\mathcal{H}_{\mathrm{B} H, K}\right|}:=\int_{0}^{T} \int_{0}^{T}|\varphi(u)||\varphi(v)| \frac{\partial^{2} \mathrm{R}_{\mathrm{H}, \mathrm{~K}}(u, v)}{\partial u \partial v} \mathrm{~d} u d v<\infty, \tag{2}
\end{equation*}
$$

such that

$$
\frac{\partial^{2} R_{H, K}(u, v)}{\partial u \partial v}=\alpha_{H, K}\left(t^{2 \mathrm{H}}+s^{2 \mathrm{H}}\right)^{\mathrm{K}-2}(\mathrm{ts})^{2 \mathrm{H}-1}+\beta_{\mathrm{H}, \mathrm{~K}}|\mathrm{t}-\mathrm{s}|^{2 \mathrm{HK}-2}
$$

where

$$
\alpha_{H, K}=2^{-K+2} \mathrm{H}^{2} \mathrm{~K}(\mathrm{~K}-1) \quad \text { and } \quad \beta_{\mathrm{H}, \mathrm{~K}}=2^{-\mathrm{K}+1} \mathrm{HK}(2 \mathrm{HK}-1) .
$$

Note that, if $\varphi, \psi \in\left|\mathcal{H}_{\mathrm{B} \boldsymbol{\beta}, \mathrm{K}}\right|$, then their scalar product in $\mathcal{H}_{\mathrm{B} \boldsymbol{\beta}, \mathrm{K}}$ is given by

$$
\langle\varphi, \psi\rangle_{\mathcal{H}_{\mathrm{B}, \mathrm{~K}}}=\int_{0}^{T} \int_{0}^{T} \varphi(u) \psi(v) \frac{\partial^{2} \mathrm{R}_{\mathrm{H}, \mathrm{~K}}(u, v)}{\partial u \partial v} \mathrm{~d} u \mathrm{~d} v .
$$

For $\varphi, \psi \in\left|\mathcal{H}_{\mathrm{B}^{\mathrm{H}, \mathrm{K}}}\right|$, we have

$$
\mathbb{E}\left(\int_{0}^{T} \varphi(u) \mathrm{dB}_{u}^{\mathrm{H}, \mathrm{~K}}\right)=0, \mathbb{E}\left(\int_{0}^{T} \varphi(u) \mathrm{dB}_{\mathfrak{u}}^{\mathrm{H}, \mathrm{~K}} \int_{0}^{\mathrm{T}} \psi(v) \mathrm{dB}_{v}^{\mathrm{H}, \mathrm{~K}}\right)=\langle\varphi, \psi\rangle_{\mathcal{H}_{\mathrm{B} H, K}} .
$$

It is worth being pointed out that the canonical Hilbert space $\mathcal{H}_{\mathrm{B}} \mathrm{H}, \mathrm{K}$ associated with $B^{H, K}$ satisfies:

$$
\begin{equation*}
\mathrm{L}^{2}([0, \mathrm{~T}]) \subset \mathrm{L}^{1 / \mathrm{HK}}([0, \mathrm{~T}]) \subset\left|\mathcal{H}_{\mathrm{B}, \mathrm{~K}}\right| \subset \mathcal{H}_{\mathrm{BH}, \mathrm{~K}}, \tag{3}
\end{equation*}
$$

where $\mathrm{H} \in(0,1), \mathrm{K} \in(0,1]$ and $\operatorname{HK} \in(1 / 2,1)$.

## 3 Preliminaries

Let $\left\{X_{t}, 0 \leq t \leq T\right\}$ be a process governed by the following equation:

$$
\begin{equation*}
d X_{t}=S\left(X_{t}\right) d t+\varepsilon d B_{t}^{H, K}, \quad X_{0}=x_{0}, 0 \leq t \leq T \tag{4}
\end{equation*}
$$

where $\varepsilon>0, \mathrm{~B}_{\mathrm{t}}^{\mathrm{H}, \mathrm{K}}$ a bifractional Brownian motion, and $\mathrm{S}($.$) is an unknown$ function. We suppose that $\chi_{t}$ is a solution of the following equation

$$
\begin{equation*}
\frac{\mathrm{d} x_{\mathrm{t}}}{\mathrm{dt}}=\mathrm{S}\left(\mathrm{x}_{\mathrm{t}}\right), \mathrm{x}_{0}, 0 \leq \mathrm{t} \leq \mathrm{T} \tag{5}
\end{equation*}
$$

We also suppose that the function $S: \mathbb{R} \longrightarrow \mathbb{R}$ satisfies the following assumptions:
(A1) There exists $L>0$ such that

$$
\begin{equation*}
|S(x)-S(y)| \leq L|x-y|, \quad x, y \in \mathbb{R} \tag{6}
\end{equation*}
$$

(A2) There exists $M>0$ such that

$$
|S(x)| \leq M(1+|x|), x \in \mathbb{R}
$$

Then, the stochastic differential equation (4) has a unique solution $\left\{X_{t}, 0 \leq t \leq T\right\}$.
(A3) Assume that the function $S(x)$ is bounded by a constant $C$.

Since the function $\chi_{\mathrm{t}}$ satisfies (5), it follows that

$$
\left|S\left(x_{\mathrm{t}}\right)-\mathrm{S}\left(x_{s}\right)\right| \leq \mathrm{L}\left|x_{\mathrm{t}}-x_{s}\right|=\mathrm{L}\left|\int_{s}^{\mathrm{t}} \mathrm{~S}\left(x_{\mathrm{r}}\right) \mathrm{dr}\right| \leq \mathrm{LC}|\mathrm{t}-\mathrm{s}|, \mathrm{t}, \mathrm{~s} \in[0, \mathrm{~T}]
$$

Let us define $\Sigma_{0}(L)$ as the class of all functions $S(x)$ satisfying the assumption (A1) and uniformly bounded by the same constant C. Further, we denote by $\Sigma_{k}(L)$ the class of all function $S(x)$ which are uniformly bounded by the same constant C and which are k-times differentiable with respect to x satisfying the following condition

$$
\begin{equation*}
\left|S^{k}(x)-S^{k}(y)\right| \leq L|x-y|, x, y \in \mathbb{R} \tag{7}
\end{equation*}
$$

where $S^{k}(x)$ is the k-th derivative of $S(x)$.
Lemma 1 Assume that hypothesis (A1) is verified. Let $X_{t}$ and $X_{t}$ be the solutions of the equations (4) and (5) respectively. Then, we have

$$
\begin{equation*}
\sup _{0 \leq t \leq T} \mathbb{E}\left(X_{t}-x_{t}\right)^{2} \leq e^{2 L T} \varepsilon^{2} T^{2 H K} \tag{8}
\end{equation*}
$$

## Proof of the Lemma 1

By (4) and (5), we have

$$
X_{t}=x_{0}+\int_{0}^{t} S\left(X_{r}\right) d r+\varepsilon B_{t}^{H, K},
$$

and

$$
x_{\mathrm{t}}=x_{0}+\int_{0}^{\mathrm{t}} S\left(x_{\mathrm{r}}\right) \mathrm{dr}
$$

This implies

$$
X_{t}-x_{t}=\int_{0}^{t}\left(S\left(X_{r}\right)-S\left(x_{r}\right)\right) d r+\varepsilon B_{t}^{H, K}
$$

Thus

$$
\begin{align*}
\left|X_{t}-x_{t}\right| & \leq \int_{0}^{t}\left|S\left(X_{r}\right)-S\left(x_{r}\right)\right| d r+\varepsilon\left|B_{t}^{\mathrm{H}, \mathrm{~K}}\right|  \tag{9}\\
& \leq L \int_{0}^{\mathrm{t}}\left|X_{r}-x_{r}\right| d r+\varepsilon\left|B_{t}^{H, K}\right| .
\end{align*}
$$

Putting $u_{t}=\left|X_{t}-x_{t}\right|$, we have

$$
u_{t} \leq \int_{0}^{t} u_{r} d r+\varepsilon\left|B_{t}^{H, K}\right|
$$

By using Grönwall's inequality, we obtain

$$
\left|X_{\mathrm{t}}-x_{\mathrm{t}}\right| \leq e^{\mathrm{Lt}} \varepsilon\left|B_{\mathrm{t}}^{\mathrm{H}, \mathrm{~K}}\right| .
$$

Then, since $\mathbb{E}\left(B_{t}^{\mathrm{H}, \mathrm{K}}\right)^{2}=\mathrm{t}^{2 \mathrm{HK}}$, we have

$$
\mathbb{E}\left|X_{t}-x_{t}\right|^{2} \leq e^{2 L t} \varepsilon^{2} t^{2 H K} .
$$

Finally, we find

$$
\sup _{0 \leq t \leq T} \mathbb{E}\left(X_{t}-x_{t}\right)^{2}<e^{2 L T} \varepsilon^{2} T^{2 H K} .
$$

## 4 Main results

The main goal of this work is to build an estimator of the trend function $S_{\mathrm{t}}$ in the model described by stochastic differential equation (4) using the method developed by Kutoyants [7]. Then, we study its asymptotic properties as $\varepsilon \longrightarrow 0$.
For all $t \in[0, T]$, the kernel estimator $\widehat{S}_{t}$ of $S_{t}$ is given by

$$
\begin{equation*}
\hat{S}_{t}=\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d X_{\tau}, \tag{10}
\end{equation*}
$$

where $G(u)$ is a bounded function with finite support $[A, B]$ satisfying the following hypotheses:
(H1) $G(u)=0$ for $u<A$ and $u>B$ and $\int_{A}^{B} G(u) d u=1$,
(H2) $\int_{-\infty}^{+\infty} G^{2}(u) d u<\infty$,
(H3) $\int_{-\infty}^{+\infty} u^{2(k+1)} \mathrm{G}^{2}(u) d u<\infty$,
(H4) $\int_{-\infty}^{+\infty}|G(u)|^{\frac{1}{H K}} d u<\infty$,
Further, we suppose that the normalizing function $\phi_{\varepsilon}$ satisfies:
(H5) $\phi_{\varepsilon} \longrightarrow 0$ and $\varepsilon^{2} \phi_{\varepsilon}^{-1} \longrightarrow 0$ as $\varepsilon \longrightarrow 0$.
The following theorem gives the uniform convergence of the estimator $\widehat{S}_{\mathrm{t}}$.
Theorem 1 Suppose that the assumptions (A1)-(A3) and (H1)-(H5) hold true. Further, suppose that the trend function $\mathrm{S}(\mathrm{x})$ belongs to $\Sigma_{0}(\mathrm{~L})$. Then, for any
$0<\mathrm{c} \leq \mathrm{d}<\mathrm{T}$ and $\mathrm{HK} \in(1 / 2,1)$, the estimator $\widehat{\mathrm{S}}_{\mathrm{t}}$ is uniformly consistent, that is,

$$
\begin{equation*}
\lim _{\varepsilon \longrightarrow 0} \sup _{S(x) \in \Sigma_{0}(L)} \sup _{c \leq t \leq d} \mathbb{E}_{S}\left(\left|\widehat{S}_{t}-S\left(x_{t}\right)\right|^{2}\right)=0 \tag{11}
\end{equation*}
$$

The following additional assumptions are useful for the rest of the theoretical study. Assume that
(H6) $\int_{-\infty}^{+\infty} u^{j} G(u) d u=0$ for $j=1,2, \ldots, k$,
(H7) $\int_{-\infty}^{+\infty} u^{k+1} G(u) d u<\infty$ and $\int_{-\infty}^{+\infty} u^{2(k+2)} G^{2}(u) d u<\infty$.
The rate of convergence of the estimator $\widehat{S}_{t}$ is established in the following theorem.

Theorem 2 Suppose that the function $\mathrm{S}(\mathrm{x}) \in \Sigma_{\mathrm{k}}(\mathrm{L}), \mathrm{HK} \in(1 / 2,1)$ and $\phi_{\varepsilon}=$ $\varepsilon^{\frac{1}{k-H K+2}}$. Then, under the hypotheses (A1)-(A3) and (H1)-(H7), we have

$$
\begin{equation*}
\limsup _{\varepsilon \longrightarrow 0} \sup _{S(x) \in \Sigma_{k}(L)} \sup _{c \leq t \leq d} \mathbb{E}_{S}\left(\left|\widehat{S}_{t}-S\left(x_{t}\right)\right|^{2}\right) \varepsilon^{\frac{-2(k+1)}{k-H K+2}}<\infty . \tag{12}
\end{equation*}
$$

Finally, the following theorem presents the asymptotic normality of the kernel type estimator $\hat{S}_{t}$ of $S\left(x_{t}\right)$.

Theorem 3 Suppose that the function $\mathrm{S}(\mathrm{x}) \in \Sigma_{\mathrm{k}+1}(\mathrm{~L}), \mathrm{HK} \in(1 / 2,1)$ and $\phi_{\varepsilon}=\varepsilon^{\frac{1}{\mathrm{k}-\mathrm{HK}+2}}$. Then, under the hypotheses (A1)-(A3) and (H1)-(H7), we have

$$
\varepsilon^{\frac{-(\mathrm{k}+1)}{\mathrm{k}-\mathrm{HK}+2}}\left(\hat{S}_{\mathrm{t}}-\mathrm{S}\left(\chi_{\mathrm{t}}\right)\right) \xrightarrow{\mathcal{D}} \mathcal{N}\left(\mathrm{m}, \sigma_{\mathrm{H}, \mathrm{~K}}^{2}\right), \text { as } \varepsilon \longrightarrow 0,
$$

where

$$
m=\frac{S^{k+1}\left(x_{t}\right)}{(k+1)!} \int_{-\infty}^{+\infty} G(u) u^{k+1} d u
$$

and

$$
\begin{aligned}
\sigma_{\mathrm{H}, \mathrm{~K}}^{2}= & \int_{-\infty}^{+\infty} \int_{-\infty}^{+\infty} \mathrm{G}(u) \mathrm{G}(v)\left[\alpha_{\mathrm{H}, \mathrm{~K}}\left(u^{2 \mathrm{H}}+v^{2 \mathrm{H}}\right)^{\mathrm{K}-2}(u v)^{2 \mathrm{H}-1}\right. \\
& \left.+\beta_{\mathrm{H}, \mathrm{~K}}|u-v|^{2 \mathrm{HK}-2}\right] d u d v
\end{aligned}
$$

with

$$
\alpha_{\mathrm{H}, \mathrm{~K}}=2^{-\mathrm{K}+2} \mathrm{H}^{2} \mathrm{~K}(\mathrm{~K}-1) \quad \text { and } \quad \beta_{\mathrm{H}, \mathrm{~K}}=2^{-\mathrm{K}+1} \mathrm{HK}(2 \mathrm{HK}-1)
$$

## 5 Numerical example

The main objective of this part is to conduct a numerical study to illustrate our theoretical result. We compare our kernel estimator for stochastic differential equations driven by a bifractional Brownian motion to the kernel estimator for stochastic differential equations driven by fractional Brownian motion given in Mishra and Prakasa Rao [10]. We compare numerically the variance $\sigma_{\mathrm{H}, \mathrm{K}}^{2}$ of our estimator to $\sigma_{\mathrm{H}}^{2}$.

Consider a function $G$ which satisfies hypotheses (H1)-(H7):

$$
G(t)=\frac{15}{128}\left(63 t^{4}+70 t^{2}+15\right),|t| \leq 1 .
$$

- The variance of the kernel estimator for stochastic differential equations driven by fractional Brownian motion given in Mishra and Prakasa Rao [10] is given as:

For all $\mathrm{H} \in(1 / 2,1)$,

$$
\sigma_{\mathrm{H}}^{2}=\mathrm{H}(2 \mathrm{H}-1) \int_{-\infty}^{+\infty} \int_{-\infty}^{+\infty} \mathrm{G}(u) \mathrm{G}(v)|u-v|^{2 \mathrm{H}-2} \mathrm{~d} u \mathrm{~d} v
$$

- Using the result given in Theorem 3, the variance of our estimator is obtained as:

For all $\mathrm{H} \in(0,1), \mathrm{K} \in(0,1]$ and $\mathrm{HK} \in(1 / 2,1)$, we have

$$
\begin{aligned}
\sigma_{\mathrm{H}, \mathrm{~K}}^{2}= & \int_{-\infty}^{+\infty} \int_{-\infty}^{+\infty} \mathrm{G}(\mathrm{u}) \mathrm{G}(v)\left[\alpha_{\mathrm{H}, \mathrm{~K}}\left(u^{2 \mathrm{H}}+v^{2 \mathrm{H}}\right)^{\mathrm{K}-2}(u v)^{2 \mathrm{H}-1}\right. \\
& \left.+\beta_{\mathrm{H}, \mathrm{~K}}|u-v|^{2 \mathrm{HK}-2}\right] \mathrm{d} u d v,
\end{aligned}
$$

where

$$
\alpha_{H, K}=2^{-K+2} \mathrm{H}^{2} \mathrm{~K}(\mathrm{~K}-1) \quad \text { and } \quad \beta_{H, K}=2^{-K+1} \mathrm{HK}(2 \mathrm{HK}-1) .
$$

Next, we compute the variances, the results are presented in the following Tables

Table 1: The variance values $\sigma_{\mathrm{H}}^{2}$.

| H | 0.7 | 0.75 | 0.8 | 0.85 | 0.9 | 0.95 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\sigma_{\mathrm{H}}^{2}$ | 1.1567 | 1.1900 | 1.1830 | 1.1506 | 1.1025 | 1.0452 |

Table 2: The variance values $\sigma_{\mathrm{H}, \mathrm{K}}^{2}$.

| $\mathrm{K} \backslash \mathrm{H}$ | 0.7 | 0.75 | 0.8 | 0.85 | 0.9 | 0.95 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0.75 | 0.6006 | 0.9458 | 1.1647 | 1.2965 | 1.3709 | 1.4091 |
| 0.8 | 0.8733 | 1.1230 | 1.2696 | 1.3462 | 1.3774 | 1.3801 |
| 0.85 | 1.0362 | 1.2107 | 1.3019 | 1.3376 | 1.3378 | 1.3159 |
| 0.9 | 1.1227 | 1.2382 | 1.2873 | 1.2930 | 1.2712 | 1.2326 |
| 0.95 | 1.1570 | 1.2264 | 1.2437 | 1.2274 | 1.1901 | 1.1402 |
| 1 | 1.1567 | 1.1900 | 1.1830 | 1.1506 | 1.1025 | 1.0452 |

From the obtained results in Tables 1 and 2, we clearly see that the variance of our estimator is less than that of the kernel estimator for stochastic differential equations driven by fractional Brownian motion. We can conclude that our kernel estimator for stochastic differential equations driven by a bifractional Brownian motion is better than that given in Mishra and Prakasa Rao [10].

## 6 Proof of Theorems

### 6.1 Proof of Theorem 1

From (4) and (10), we can see that

$$
\begin{aligned}
\hat{S}_{t}-S\left(x_{t}\right)= & \frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d X_{\tau}-S\left(x_{t}\right) \\
= & \frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right)\left(S\left(X_{\tau}\right) d \tau+\varepsilon d B_{\tau}^{H, K}\right)-S\left(x_{t}\right) \\
= & \frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right)\left(S\left(X_{\tau}\right)-S\left(x_{\tau}\right)\right) d \tau \\
& +\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) S\left(x_{\tau}\right) d \tau-S\left(x_{t}\right) \\
& +\frac{\varepsilon}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d B_{\tau}^{H, K}
\end{aligned}
$$

Using the inequality $(\alpha+\beta+\gamma)^{2} \leq 3 \alpha^{2}+3 \beta^{2}+3 \gamma^{2}$, it yields

$$
\begin{align*}
\mathbb{E}_{S}\left[\hat{S}_{t}-S\left(x_{t}\right)\right]^{2} \leq & 3 \mathbb{E}_{S}\left[\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right)\left(S\left(X_{\tau}\right)-S\left(x_{\tau}\right)\right) d \tau\right]^{2} \\
& +3 \mathbb{E}_{S}\left[\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) S\left(x_{\tau}\right) d \tau-S\left(x_{t}\right)\right]^{2}  \tag{13}\\
& +3 \mathbb{E}_{S}\left[\frac{\varepsilon}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) \mathrm{dB}_{\tau}^{\mathrm{H}, \mathrm{~K}}\right]^{2} \\
\leq & I_{1}+I_{2}+I_{3}
\end{align*}
$$

- Concerning $\mathrm{I}_{1}$. Via inequalities (6) and (8) and hypotheses (H1)-(H2), we get

$$
\begin{align*}
I_{1} & =3 \mathbb{E}_{S}\left[\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right)\left(S\left(X_{\tau}\right)-S\left(x_{\tau}\right)\right) d \tau\right]^{2} \\
& =3 \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G(u)\left(S\left(X_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t+\phi_{\varepsilon} u}\right)\right) d u\right]^{2} \\
& \leq 3(B-A) \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G^{2}(u)\left(S\left(X_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t+\phi_{\varepsilon} u}\right)\right)^{2} d u\right] \\
& \leq 3(B-A) L^{2} \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G^{2}(u)\left(X_{t+\phi_{\varepsilon} u}-x_{t+\phi_{\varepsilon} u}\right)^{2} d u\right]  \tag{14}\\
& \leq 3(B-A) L^{2}\left[\int_{-\infty}^{+\infty} G^{2}(u) \sup _{0 \leq t+\phi_{\varepsilon} u \leq T} \mathbb{E}_{S}\left(X_{t+\phi_{\varepsilon} u}-x_{t+\phi_{\varepsilon} u}\right)^{2} d u\right] \\
& \leq 3(B-A) L^{2} e^{2 L T} T^{2 H K} \varepsilon^{2} \\
& \leq C_{1} \varepsilon^{2},
\end{align*}
$$

where $C_{1}$ is a positive constant depending on $T, L, H, K$, and $(B-A)$.

- Concerning $I_{2}$. Let

$$
\begin{align*}
I_{2} & =3 \mathbb{E}_{S}\left[\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) S\left(x_{\tau}\right) d \tau-S\left(x_{t}\right)\right]^{2} \\
& =3 \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G(u) S\left(x_{t+\phi_{\varepsilon} u}\right) d u-S\left(x_{t}\right)\right]^{2}  \tag{15}\\
& =3 \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G(u)\left(S\left(x_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t}\right)\right) d u\right]^{2}
\end{align*}
$$

Next, by using hypotheses (A3) and (H3), we have

$$
\begin{aligned}
\mathrm{I}_{2} & \leq 3 \mathrm{~L}^{2} \mathrm{C}_{2}^{2} \mathbb{E}_{\mathrm{S}}\left[\int_{-\infty}^{+\infty} \mathrm{G}(u)\left(\phi_{\varepsilon} u\right) \mathrm{du}\right]^{2} \\
& \leq 3(\mathrm{~B}-A) \mathrm{L}^{2} \mathrm{C}_{2}^{2}\left[\int_{-\infty}^{+\infty} \mathrm{G}^{2}(u) u^{2} d u\right] \phi_{\varepsilon}^{2} \\
& \leq \mathrm{C}_{3} \phi_{\varepsilon}^{2}
\end{aligned}
$$

where $C_{3}$ is a positive constant depending on $L$ and $(B-A)$.

- Concerning $I_{3}$. Since $H K \in(1 / 2,1)$, we have

$$
\begin{aligned}
\mathrm{I}_{3} & =3 \mathbb{E}_{\mathrm{S}}\left[\frac{\varepsilon}{\phi_{\varepsilon}} \int_{0}^{T} \mathrm{G}\left(\frac{\tau-\mathrm{t}}{\phi_{\varepsilon}}\right) \mathrm{dB}_{\tau}^{\mathrm{H}, \mathrm{~K}}\right]^{2} \\
& =3 \frac{\varepsilon^{2}}{\phi_{\varepsilon}^{2}} \mathbb{E}_{\mathrm{S}}\left[\int_{0}^{T} \mathrm{G}\left(\frac{\tau-\mathrm{t}}{\phi_{\varepsilon}}\right) \mathrm{dB}_{\tau}^{\mathrm{H}, \mathrm{~K}}\right]^{2} \\
& \leq 3 \frac{\varepsilon^{2}}{\phi_{\varepsilon}^{2}}\left[\mathrm { C } ( 2 , \mathrm { HK } ) \left(\int_{0}^{\mathrm{T}} \left\lvert\, \mathrm{G}\left(\frac{\tau-\mathrm{t}}{\phi_{\varepsilon}}\right)^{\left.\left.\frac{1}{\mathrm{HK}} \mathrm{~d} \tau\right)^{2 \mathrm{HK}}\right]}\right.\right.\right. \\
& \leq \mathrm{C}_{4} \frac{\varepsilon^{2}}{\phi_{\varepsilon}^{2}}\left[\phi_{\varepsilon}^{2 \mathrm{HK}}\left(\int_{-\infty}^{+\infty}|\mathrm{G}(u)|^{\frac{1}{\mathrm{HK}}} \mathrm{du}\right)^{2 \mathrm{HK}}\right] \\
& \leq \mathrm{C}_{5} \frac{\varepsilon^{2}}{\phi_{\varepsilon}} \phi_{\varepsilon}^{2 \mathrm{HK}-1} \quad(\text { using hypothesis }(\mathrm{H} 4))
\end{aligned}
$$

where $\mathrm{C}_{5}$ is a positive constant depending on H and K .
Combining (13)-(16), we have

$$
\sup _{S(x) \in \Sigma_{0}(\mathrm{~L})} \sup _{c \leq t \leq \mathrm{d}} \mathbb{E}_{S}\left[\hat{S}_{t}-S\left(x_{t}\right)\right]^{2} \leq \mathrm{C}_{6}\left(\varepsilon^{2}+\phi_{\varepsilon}^{2}+\frac{\varepsilon^{2}}{\phi_{\varepsilon}} \phi_{\varepsilon}^{2 \mathrm{HK}-1}\right)
$$

Finally, under the assumption (H5), we obtain

$$
\lim _{\varepsilon \longrightarrow 0} \sup _{S(x) \in \Sigma_{0}(L)} \sup _{c \leq t \leq d} \mathbb{E}_{S}\left[\hat{S}_{t}-S\left(x_{t}\right)\right]^{2}=0
$$

### 6.2 Proof of Theorem 2

Using the Taylor formula, we get

$$
S\left(x_{t}\right)=S\left(x_{t_{0}}\right)+\sum_{j=1}^{k} S^{j}\left(x_{t_{0}}\right) \frac{\left(t-t_{0}\right)^{j}}{j!}
$$

$$
+\left(S^{k}\left(x_{t+\lambda\left(t-t_{0}\right)}\right)-S^{k}\left(x_{t_{0}}\right)\right) \frac{\left(t-t_{0}\right)^{k}}{k!}, \lambda \in(0,1)
$$

and

$$
\begin{aligned}
S\left(x_{t+\phi_{\varepsilon} u}\right)= & S\left(x_{t}\right)+\sum_{j=1}^{k} S^{j}\left(x_{t}\right) \frac{\left(\phi_{\varepsilon} u\right)^{j}}{j!} \\
& +\left(S^{k}\left(x_{t+\lambda\left(\phi_{\varepsilon} u\right)}\right)-S^{k}\left(x_{t}\right)\right) \frac{\left(\phi_{\varepsilon} u\right)^{k}}{k!}, \lambda \in(0,1) .
\end{aligned}
$$

Then, by substituting this expression in $\mathrm{I}_{2}$, using inequality (7) and assumptions (H6)-(H7), we obtain

$$
\begin{align*}
I_{2} & =3 \mathbb{E}_{S}\left[\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) S\left(x_{\tau}\right) d \tau-S\left(x_{t}\right)\right]^{2} \\
& =3 \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G(u) S\left(x_{t+\phi_{\varepsilon} u}\right) d u-S\left(x_{t}\right)\right]^{2} \\
& =3 \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G(u)\left(S\left(x_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t}\right)\right) d u\right]^{2} \\
& =3 \mathbb{E}_{S}\left[\int_{-\infty}^{+\infty} G(u)\left(\sum_{j=1}^{k} S^{j}\left(x_{t}\right) \frac{\left(\phi_{\varepsilon} u\right)^{j}}{j!}+\left(S^{k}\left(x_{t+\lambda\left(\phi_{\varepsilon} u\right)}\right)-S^{k}\left(x_{t}\right)\right) \frac{\left(\phi_{\varepsilon} u\right)^{k}}{k!}\right) d u\right]^{2} \\
& =3 \mathbb{E}_{S}\left[\frac{\phi_{\varepsilon}^{k}}{k!} \int_{-\infty}^{+\infty} G(u) u^{k}\left(S^{k}\left(x_{t+\lambda\left(\phi_{\varepsilon} u\right)}\right)-S^{k}\left(x_{t}\right)\right) d u\right]^{2}(b y u \operatorname{sing}(H 6)) \\
& \leq 3 C_{7}^{2} L^{2}\left[\frac{\phi_{\varepsilon}^{k+1}}{k!} \int_{-\infty}^{+\infty} G(u) u^{k+1} d u\right]^{2} \\
& \leq 3 C_{7}^{2} L^{2}(B-A) \frac{\phi_{\varepsilon}^{2(k+1)}}{(k!)^{2}}\left[\int_{-\infty}^{+\infty} G^{2}(u) u^{2(k+1)} d u\right] \\
& \leq C_{8} \phi_{\varepsilon}^{2(k+1)}, \tag{17}
\end{align*}
$$

where $C_{8}$ is a positive constant depending on $L$ and $(B-A)$.
Next, from (14), (16), and (17), we find

$$
\sup _{S(x) \in \Sigma_{k}(L)} \sup _{c \leq t \leq d} \mathbb{E}_{S}\left|\hat{S}_{t}-S\left(x_{t}\right)\right|^{2} \leq C_{9}\left(\varepsilon^{2} \phi_{\varepsilon}^{2 H K-2}+\phi_{\varepsilon}^{2(k+1)}+\varepsilon^{2}\right)
$$

Putting $\phi_{\varepsilon}=\varepsilon^{\frac{1}{k-H K+2}}$, it yields

$$
\limsup _{\varepsilon \longrightarrow 0} \sup _{S(x) \in \Sigma_{k}(L)} \sup _{c \leq t \leq d} \mathbb{E}_{S}\left(\left|\hat{S}_{t}-S\left(x_{t}\right)\right|^{2}\right) \varepsilon^{\frac{-2(k+1)}{k-H K+2}}<\infty .
$$

This completes the proof of Theorem 2.

### 6.3 Proof of Theorem 3

From (4) and (10), we can see that

$$
\begin{aligned}
& \varepsilon^{\frac{-(k+1)}{k-H K+2}}\left(\hat{S}_{t}-S\left(X_{t}\right)\right)=\varepsilon^{\frac{-(k+1)}{k-H K+2}}\left[\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right)\left(S\left(X_{\tau}\right)-S\left(X_{\tau}\right)\right) d \tau\right. \\
& \left.+\frac{1}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) S\left(x_{\tau}\right) d \tau-S\left(x_{t}\right)+\frac{\varepsilon}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d B_{\tau}^{H, K}\right] .
\end{aligned}
$$

Therefore

$$
\begin{aligned}
\varepsilon^{\frac{-(k+1)}{k-H K+2}}\left(\widehat{S}_{t}\right. & \left.-S\left(x_{t}\right)\right)=\varepsilon^{\frac{-(k+1)}{k-H K+2}}\left[\int_{-\infty}^{+\infty} G(u)\left(S\left(X_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t+\phi_{\varepsilon} u}\right)\right) d u\right. \\
& \left.+\int_{-\infty}^{+\infty} G(u)\left(S\left(x_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t}\right)\right) d u+\frac{\varepsilon}{\phi_{\varepsilon}} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d S_{\tau}^{H}\right]
\end{aligned}
$$

Thus

$$
\varepsilon^{\frac{-(k+1)}{k-H K+2}}\left(\hat{S}_{t}-S\left(x_{t}\right)\right)=r_{1}(t)+r_{2}(t)+\eta_{\varepsilon}(t)
$$

Hence, by Slutsky's Theorem, it suffices to show the following three claims:

$$
\begin{align*}
& \mathrm{r}_{1}(\mathrm{t}) \rightarrow 0, \text { as } \varepsilon \rightarrow 0 \text { in probability. }  \tag{18}\\
& \mathrm{r}_{2}(\mathrm{t}) \rightarrow \mathrm{m}, \text { as } \varepsilon \rightarrow 0 \text { in probability. } \tag{19}
\end{align*}
$$

and

$$
\begin{equation*}
\eta_{\varepsilon}(t) \rightarrow \mathcal{N}\left(0, \sigma_{H, K}^{2}\right), \text { as } \varepsilon \rightarrow 0 \text { in distribution. } \tag{20}
\end{equation*}
$$

Proof of (18).
Let

$$
r_{1}(t)=\varepsilon^{\frac{-(k+1)}{k-H K+2}} \int_{-\infty}^{+\infty} G(u)\left(S\left(X_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t+\phi_{\varepsilon} u}\right)\right) d u
$$

By applying the inequality (14), we have

$$
0 \leq \mathbb{E}\left[r_{1}^{2}(t)\right] \leq \varepsilon^{\frac{-2(k+1)}{k-H K+2}} I_{1} \leq C_{10} \varepsilon^{\frac{2(1-H K)}{k-H K+2}}
$$

Therefore, using the Bienaymé-Tchebychev's inequality, as $\varepsilon \longrightarrow 0$, we obtain, for all $\alpha>0$

$$
P\left(\left|r_{1}(t)\right|>\alpha\right) \leq \frac{\mathbb{E}\left[r_{1}^{2}(t)\right]}{\alpha^{2}} \leq \frac{C_{10} \varepsilon^{\frac{2(1-H K)}{k-H K+2}}}{\alpha^{2}} \longrightarrow 0
$$

## Proof of (19).

Let

$$
r_{2}(t)=\varepsilon^{\frac{-(k+1)}{k-H K+2}} \int_{-\infty}^{+\infty} G(u)\left(S\left(x_{t+\phi_{\varepsilon} u}\right)-S\left(x_{t}\right)\right) d u
$$

By taking any $t, u \in[0, T]$ and $b(x) \in \Sigma_{k+1}(L)$, via the Taylor expansion, we get

$$
\begin{aligned}
S\left(x_{t+\phi_{\varepsilon} u}\right)= & S\left(x_{t}\right)+\sum_{j=1}^{k} S^{j}\left(x_{t}\right) \frac{\left(\phi_{\varepsilon} u\right)^{j}}{j!}+\frac{S^{k+1}\left(x_{t}\right)}{(k+1)!}\left(\phi_{\varepsilon} u\right)^{k+1} \\
& +\left(S^{k+1}\left(x_{t+\lambda\left(\phi_{\varepsilon} u\right)}\right)-S^{k+1}\left(x_{t}\right)\right) \frac{\left(\phi_{\varepsilon} u\right)^{k+1}}{(k+1)!}, \lambda \in(0,1)
\end{aligned}
$$

Making use of the conditions (H6), (H7), and choosing $\phi_{\varepsilon}=\varepsilon^{\frac{1}{k-H K+2}}$, we obtain

$$
\begin{aligned}
\mathbb{E}\left[r_{2}(t)-m\right]^{2} & =\mathbb{E}\left[\int_{-\infty}^{+\infty} G(u)\left(S^{k+1}\left(x_{t+\lambda\left(\phi_{\varepsilon} u\right)}\right)-S^{k+1}\left(x_{t}\right)\right) \frac{(u)^{k+1}}{(k+1)!} d u\right]^{2} \\
& \leq C_{11} L^{2} C^{2}\left(\int_{-\infty}^{+\infty} G(u) u^{k+2} \frac{\phi_{\varepsilon}}{(k+1)!} d u\right)^{2} \\
& \leq C_{12}\left(\int_{-\infty}^{+\infty} G^{2}(u) u^{2(k+2)} d u\right) \phi_{\varepsilon}^{2} \\
& \leq C_{13} \phi_{\varepsilon}^{2},
\end{aligned}
$$

where $C_{13}$ is a positive constant which depends on $L$ and $k$, and

$$
m=\frac{S^{k+1}\left(x_{t}\right)}{(k+1)!} \int_{-\infty}^{+\infty} G(u) u^{k+1} d u
$$

Therefore,

$$
\mathbb{E}\left[r_{2}(t)-m\right]^{2} \longrightarrow 0 \text { as } \varepsilon \longrightarrow 0
$$

Then

$$
r_{2}(t) \xrightarrow{\mathbb{P}} m .
$$

Proof of (20).
Let

$$
\begin{equation*}
\eta_{\varepsilon}(t)=\varepsilon^{\frac{-(k+1)}{k-H K+2}} \varepsilon \phi_{\varepsilon}^{-1} \int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d B_{\tau}^{H, K} \tag{21}
\end{equation*}
$$

In fact, we have to evaluate the variance of (21). To this end, let

$$
\mathbb{E}\left[\eta_{\varepsilon}(t)\right]^{2}=\left(\varepsilon^{\frac{1-H K}{k-H K+2}} \phi_{\varepsilon}^{-1}\right)^{2} \mathbb{E}\left(\int_{0}^{T} G\left(\frac{\tau-t}{\phi_{\varepsilon}}\right) d B_{\tau}^{\mathrm{H}, \mathrm{~K}}\right)^{2} .
$$

Moreover, using equation (2), we have

$$
\mathbb{E}\left[\eta_{\varepsilon}(t)\right]^{2}=\left(\varepsilon^{\frac{1-H K}{k-H K+2}} \phi_{\varepsilon}^{-1}\right)^{2}\left[\phi_{\varepsilon}^{2 H K} \int_{-\infty}^{+\infty} \int_{-\infty}^{+\infty} G(u) G(v) \frac{\partial^{2} R_{H, K}(u, v)}{\partial u \partial v} d u d v\right] .
$$

Then, by taking $\phi_{\varepsilon}=\varepsilon^{\frac{1}{k-H K+2}}$, we get

$$
\mathbb{E}\left[\eta_{\varepsilon}(t)\right]^{2}=\int_{-\infty}^{+\infty} \int_{-\infty}^{+\infty} G(u) G(v) \frac{\partial^{2} R_{H, K}(u, v)}{\partial u \partial v} d u d v
$$

with

$$
\frac{\partial^{2} \mathrm{R}_{\mathrm{H}, \mathrm{~K}}(u, v)}{\partial u \partial v}=\alpha_{\mathrm{H}, \mathrm{~K}}\left(u^{2 \mathrm{H}}+v^{2 \mathrm{H}}\right)^{\mathrm{K}-2}(u v)^{2 \mathrm{H}-1}+\beta_{\mathrm{H}, \mathrm{~K}}|u-v|^{2 \mathrm{HK}-2},
$$

where

$$
\alpha_{H, K}=2^{-K+2} H^{2} K(K-1) \quad \text { and } \quad \beta_{H, K}=2^{-K+1} \mathrm{HK}(2 \mathrm{HK}-1) .
$$

Finally, this last equation allows us to achieve the proof of Theorem 3.

## 7 Conclusion

This paper considered a nonparametric estimation of trend function for stochastic differential equations driven by a bifractional Brownian motion. We constructed an estimate of the trend function. Then, under some assumptions, we established the uniform consistency, the rate of convergence and the asymptotic normality of the proposed estimator. Further, an numerical example is provided. The present study has many applications in practical phenomena including telecommunications and economics.

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## References

[1] X. Bardina, K. Es-Sebaiy, An extension of bifractional Brownian motion, Communications on Stochastic Analysis, 5 (2) (2011), 333-340.
[2] K. Es-Sebaiy and C. A. Tudor, Multidimensional bifractional Brownian motion: Itô and Tanaka formulas, Stoch. Dyn., 7 (3) (2007), 365-388.
[3] C. Houdré and J. Villa, An example of infinite dimensional quasi-helix, Stochastic models (Mexico City, 2002), Contemp. Math., 336 (2003), 195201.
[4] J. P. Kahane, Hélices et quasi-hélices, Adv. Math., 7B (1981), 417-433.
[5] A. N. Kolmogorov, Wienersche Spiralen und einige andere interessante Kurven im Hilbertschen Raum. C.R. (Doklady) Acad. USSR (N.S.), 26 (1940), 115-118.
[6] I. Kruk, F. Russo, and C. A. Tudor, Wiener integrals, Malliavin calculus and covariance measure structure, J. Funct. Anal., 249 (1) (2007), 92142.
[7] Y. A. Kutoyants, Identification of dynamical systems with small noise, Springer Science \& Business Media, 300 (2012).
[8] M. Lifshits, K. Volkova, Bifractional Brownian motion: Existence and Border cases, preprint. http://arxiv.org/pdf/1502.02217.pdf., (2015).
[9] B. B. Mandelbrot, J. W. Van Ness, Fractional Brownian motions, fractional noises and applications, SIAM Rev., 10 (1968), 422-437.
[10] M. N. Mishra and B. L. S. Prakasa Rao, Nonparameteric Estimation of Trend for Stochastic Differential Equations Driven by Fractional Brownian Motion, Stat. Inference. Stoch. Process., 14 (2011), 101-109.
[11] M. N. Mishra, B. L. S. Prakasa Rao, Nonparametric Estimation of Linear Multiplier for Fractional Diffusion processes, Stochastic Analysis and Application, 29 (2011), 706-712.
[12] I. Norros, E. Valkeila, J. Virtamo, An elementary approach to a Girsanov formula and other analytical results on fractional Brownian motions, Bernoulli, 5 (4) (1999), 571-587.
[13] B. L. S. Prakasa Rao, Nonparametric estimation of trend for stochastic differential equations driven by mixed fractional Brownian motion, Stochastic Analysis and Applications, 37 (2) (2019), 271-280.
[14] F. Russo and C. Tudor, On the bifractional Brownian motion, Stoch. Process. Their Appl., 116 (5) (2006), 830-856.
[15] B. Saussereau, Nonparametric inference for fractional diffusion, Bernoulli, 20 (2) (2014), 878-918.
[16] C. A. Tudor, Y. Xiao, Sample path properties of bifractional Brownian motion, Bernoulli 13 (2007), 1023-1052.
[17] L. Yan, J. Liu, and G. Jing, Quadratic covariation and Itô formula for a bifractional Brownian motion, preprint (2008).

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# Orbital shadowing property on chain transitive sets for generic diffeomorphisms 

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#### Abstract

Let $f: M \rightarrow M$ be a diffeomorphism on a closed smooth $n(\geq 2)$ dimensional manifold $M$. We show that $C^{1}$ generically, if a diffeomorphism $f$ has the orbital shadowing property on locally maximal chain transitive sets which admits a dominated splitting then it is hyperbolic.


## 1 Introduction

Let $M$ be a closed smooth $\mathfrak{n}(n \geq 2)$-dimensional Riemannian manifold, and let $\operatorname{Diff}(M)$ be the space of diffeomorphisms of $M$ endowed with the $C^{1}$-topology. Denote by $d$ the distance on $M$ induced from a Riemannian metric $\|\cdot\|$ on the tangent bundle TM. For any $\delta>0$, a sequence $\left\{x_{i}\right\}_{i \in \mathbb{Z}}$ is called a $\delta$-pseudo orbit of $f$ if $d\left(f\left(x_{i}\right), x_{i+1}\right)<\delta$ for all $i \in \mathbb{Z}$. Let $\Lambda$ be a closed $f$-invariant set. We say that f has the shadowing property on $\Lambda$ if for any $\epsilon>0$ there is $\delta>0$ such that for any $\delta$-pseudo orbit $\left\{x_{i}\right\}_{i \in \mathbb{Z}} \subset \Lambda$ there is $y \in M$ such that $d\left(f^{i}(y), x_{i}\right)<\epsilon$ for all $i \in \mathbb{Z}$. If $\Lambda=M$ then we say that $f$ has the shadowing property. The shadowing property is very useful notion to investigate for hyperbolic structure. In fact, Robinson[22] and Sakai[24] proved that a diffeomorphism f has the $C^{1}$ robustly shadowing property if and only if it is structurally stable

[^5]diffeomorphisms. Here, we say that $f$ has the $C^{1}$ robustly shadowing property if there is a $C^{1}$ neighborhood $\mathcal{U}(f)$ of $f$ such that for any $g \in \mathcal{U}(f), g$ has the shadowing property. From the property, a general shadowing property was introduced by [20] which is called the orbital shadowing property. For the orbital shadowing property, many results published by the various view points (see $[10,13,14,15,16,17,19]$ ). We say that f has the orbital shadowing property on $\Lambda$ if for any $\epsilon>0$ there is $\delta>0$ such that for any $\delta$ pseudo orbit $\xi=\left\{x_{i}\right\}_{\in \mathbb{Z}} \subset \Lambda$ such that there is a point $y \in M$ such that
$$
\operatorname{Orb}(y) \subset B_{\epsilon}(\xi) \text { and } \xi \subset B_{\epsilon}(\operatorname{Orb}(y)) .
$$

If $\Lambda=M$ then we say that $f$ has the orbital shadowing property. Let $\Lambda$ be a closed f-invariant set. We say that $\Lambda$ is hyperbolic if the tangent bundle $\mathrm{T}_{\Lambda} \mathrm{M}$ has a Df-invariant splitting $\mathrm{E}^{s} \oplus \mathrm{E}^{\mathrm{u}}$ and there exist constants $\mathrm{C}>0$ and $0<\lambda<1$ such that

$$
\left\|\left.D_{x} f^{n}\right|_{E_{x}^{s}}\right\| \leq C \lambda^{n} \text { and }\left\|\left.D_{x} f^{-n}\right|_{E_{x}^{u}}\right\| \leq C \lambda^{n}
$$

for all $x \in \Lambda$ and $n \geq 0$. If $\Lambda=M$ then we say that $f$ is Anosov.
Pilyugin et al [20] proved that if a diffeomorphism f has the $\mathrm{C}^{1}$ robustly orbital shadowing property then it is structurally stable diffeomorphisms. Lee and Lee [10] proved that a volume preserving diffeomorphism $f$ has the $C^{1}$ robustly orbital shadowing property then it is Anosov. Moreover, we can find similar results $[12,13,14,15]$. We say that the set $\Lambda$ is transitive if there is a point $x \in \Lambda$ such that $\omega(x)=\Lambda$, where $\omega(x)$ is the omega limit set of $x$. An invariant closed set $\mathcal{C}$ is called a chain transitive if for any $\delta>0$ and $x, y \in \mathcal{C}$, there is $\delta$-pseudo orbit $\left\{x_{i}\right\}_{i=0}^{n}(n \geq 1) \subset \mathcal{C}$ such that $x_{0}=x$ and $x_{n}=y$. It is clear that the transitive set $\Lambda$ is the chain transitive set $\mathcal{C}$, but the converse is not true. We say that $\Lambda$ is locally maximal if there is a neighborhood U of $\Lambda$ such that $\Lambda=\bigcap_{\mathfrak{n} \in \mathbb{Z}} \mathrm{f}^{n}(\mathrm{U})$. For the relation between chain transitive sets and $C^{1}$ robustly shadowing theories, In [16], Lee proved that if a robustly chain transitive set with orbital shadowing then it is hyperbolic. We say that $f$ has the $C^{1}$ stably shadowing property on $\Lambda$ if there are a $C^{1}$ neighborhood $\mathcal{U}(f)$ of f and a neighborhood U of $\Lambda$ such that for any $\mathrm{g} \in \mathcal{U}(\mathrm{f}), \mathrm{g}$ has the shadowing property on $\Lambda_{g}(U)$, where $\Lambda_{g}(U)$ is the continuation of $\Lambda$. For $f \in \operatorname{Diff}(M)$, we say that a compact f-invariant set $\Lambda$ admits a dominated splitting if the tangent bundle $T_{\Lambda} M$ has a continuous Df-invariant splitting $E \oplus F$ and there exist $C>0,0<\lambda<1$ such that for all $x \in \Lambda$ and $n \geq 0$, we have

$$
\left\|\left.D f^{n}\right|_{E(x)}\right\| \cdot\left\|\left.D f^{-n}\right|_{F\left(f f^{n}(x)\right)}\right\| \leq C \lambda^{n} .
$$

In [18], Lee proved that if a diffeomorphism $f$ has the $C^{1}$ stably shadowing property on chain transitive set $\mathcal{C}$ then it admits a dominated splitting. Sakai [23] proved that a diffeomorphism $f$ has the $C^{1}$ stably shadowing property on chain transitive set $\mathcal{C}$ then it is hyperbolic.

We say that a subset $\mathcal{G} \subset \operatorname{Diff}(M)$ is residual if $\mathcal{G}$ contains the intersection of a countable family of open and dense subsets of $\operatorname{Diff}(M)$; in this case $\mathcal{G}$ is dense in $\operatorname{Diff}(M)$. A property " P " is said to be $\left(\mathrm{C}^{1}\right)$-generic if " P " holds for all diffeomorphisms which belong to some residual subset of $\operatorname{Diff}(M)$. For $C^{1}$ generic differeomorphism f, Abdenur and Díaz [3] suggested the problem : if a $\mathrm{C}^{1}$ generic diffeomorphism f has the shadowing property then is it hyperbolic?

Unfortunately, this question still is open. For the problem, there are partial results [4, 9, 11]. Ahn et al [4] proved that for $C^{1}$ generic diffeomorphism f , if f has the shadowing property on a locally maximal homoclinic class then it is hyperbolic. Lee and Wen [11] proved that for $C^{1}$ generic diffeomorphism $f$, if $f$ has the shadowing property on a locally maximal chain transitive set then it is hyperbolic. Very recently, Lee and Lee [9] proved that for $C^{1}$ generic diffeomorphism $f$, if $f$ has the shadowing property on chain recurrence classes then it is hyperbolic. From the results, we study the orbital shadowing property for $C^{1}$ generic diffeomorphisms. The following is the main theorem of the paper.

Theorem A For $\mathrm{C}^{1}$ generic f , if f has the orbital shadowing property on a locally maximal $\mathcal{C}$ which admits a dominated splitting $\mathrm{E} \oplus \mathrm{F}$ then it is hyperbolic.

## 2 Proof of Theorem A

Let $M$ be as before, and let $f \in \operatorname{Diff}(M)$. A periodic point for $f$ is a point $p \in M$ such that $f^{\pi(p)}(p)=p$, where $\pi(p)$ is the minimum period of $p$. Denote by $P(f)$ the set of all periodic points of $f$. Let $p$ be a hyperbolic periodic point of f . A point $x \in M$ is called chain recurrent if for any $\delta>0$, there is a finite $\delta$-pseudo orbit $\left\{x_{i}\right\}_{i=0}^{n}(n \geq 1)$ such that $x_{0}=x$ and $x_{n}=x$. Denote by $\mathcal{C} \mathcal{R}(f)$ the set of all chain recurrent points of $f$. We define a relation $\leadsto \rightarrow$ on $\mathcal{C} \mathcal{R}(f)$ by $x \nsim \leadsto y$ if for any $\delta>0$, there is a finite $\delta$ pseudo orbit $\left\{x_{i}\right\}_{i=0}^{n}$ such that $x_{0}=x$ and $x_{n}=y$ and a finite $\delta$ pseudo orbit $\left\{w_{i}\right\}_{i=0}^{n}$ such that $w_{0}=y$ and $w_{n}=x$. Then we know that the relation $\longleftrightarrow \gg$ is an equivalence relation on $\mathcal{C R}(f)$. the equivalence classes are called the chain recurrence classes of $f$, denote by $C_{f}$. Note that if the class $C_{f}$ has a hyperbolic periodic point $p$ then we denote as $C(p, f)$.

It is well known that if $p$ is a hyperbolic periodic point of $f$ with period $k$
then the sets

$$
\begin{aligned}
& W^{s}(p)=\left\{x \in M: f^{k n}(x) \rightarrow p \text { as } n \rightarrow \infty\right\} \text { and } \\
& \mathcal{W}^{\mathfrak{u}}(p)=\left\{x \in M: f^{-k n}(x) \rightarrow p \text { as } n \rightarrow \infty\right\}
\end{aligned}
$$

are $C^{1}$-injectively immersed submanifolds of $M$. The homoclinic class of a hyperbolic periodic point $p$ is the closure of the transverse intersection of the $W^{s}(p)$ and $W^{u}(p)$, and it is denoted by $H(p, f)$. It is clear that $H(p, f)$ is compact, transitive and invariant sets. Let $q$ be a hyperbolic periodic point of f . We say that p and q are homoclinically related, and write $\mathrm{p} \sim \mathrm{q}$ if

$$
W^{s}(p) \pitchfork W^{\mathfrak{u}}(q) \neq \emptyset \text { and } W^{\mathbf{u}}(p) \pitchfork W^{s}(q) \neq \emptyset .
$$

It is clear that if $p \sim q$ then $\operatorname{index}(p)=\operatorname{index}(q)$, that is, $\operatorname{dimW}^{s}(p)=$ $\operatorname{dimW}^{s}(\mathfrak{q})$. By the Smale's transverse homoclinic point theorem, $\mathrm{H}_{\mathrm{f}}(\mathfrak{p})$ coincides with the closure of the set of hyperbolic periodic points $q$ of $f$ such that $p \sim q$. Note that if $p$ is a hyperbolic periodic point of $f$ then there is a neighborhood $U$ of $p$ and a $C^{1}$-neighborhood $\mathcal{U}(f)$ of $f$ such that for any $g \in \mathcal{U}(f)$ there exists a unique hyperbolic periodic point $p_{g}$ of $g$ in $U$ with the same period as $p$ and $\operatorname{index}\left(p_{g}\right)=\operatorname{index}(p)$. Such a point $p_{g}$ is called the continuation of $p=p_{f}$. The following are results for $C^{1}$ generic diffeomorphisms (see [2]).

Lemma 1 There is a residual set $\mathcal{G} \subset \operatorname{Diff}(M)$ such that if $f \in \mathcal{G}$,
(a) $\mathrm{H}(\mathfrak{p}, \mathfrak{f})=\mathrm{C}(\mathfrak{p}, \mathrm{f})$, for some hyperbolic periodic point p (see [5]).
(b) A compact $\boldsymbol{f}$-invariant set $\mathcal{C}$ is chain transitive if and only if $\mathcal{C}$ is the Hausdorff limit of a sequence of periodic orbits of f (see [8]).
(c) A locally maximal transitive set $\Lambda$ is a locally maximal $\mathrm{H}(\mathrm{p}, \mathrm{f})$ for some periodic point $\mathrm{p} \in \Lambda$ (see [1]).
(d) $\mathrm{H}(\mathrm{p}, \mathrm{f})=\overline{\mathrm{W}^{\mathrm{s}}(\mathfrak{p})} \cap \overline{\mathrm{W}^{\mathbf{u}}(\mathfrak{p})}$ (see [7]).

Remark 1 Applying Pugh's closing lemma, we know that any transitive set $\Lambda$ of a $\mathrm{C}^{1}$-generic diffeomorphism f is the Hausdorff limit of a sequence of periodic orbits $\operatorname{Orb}_{\mathfrak{f}}\left(\mathfrak{p}_{\mathfrak{n}}\right)$ of f , that is, $\lim _{\mathfrak{n} \rightarrow \infty} \operatorname{Orb}_{\mathfrak{f}}\left(\mathfrak{p}_{\mathfrak{n}}\right)=\Lambda$. By Lemma 1 (b) and (c), a chain transitive set $\mathcal{C}$ is a transitive set $\Lambda$ and so, a locally maximal chain transitive set $\mathcal{C}=\mathrm{H}(\mathrm{p}, \mathrm{f})$ for some periodic point p .

Let $\Lambda$ be a closed f-invariant set. We say that $\Lambda$ is Lyapunov stable for f if for any open neighborhood U of $\Lambda$ there is a neighborhood $\mathrm{V} \subset \mathrm{U}$ such that
$f^{j}(\mathrm{~V}) \subset \mathrm{U}$ for all $\mathfrak{j} \in \mathbb{N}$. We say that the closed set is bi-Lyapunov stable if it is Lyapunov stable for $f$ and $f^{-1}$. Potrie [21, Theorem 1.1] proved that $C^{1}$ generically, if a homoclinic class $\mathrm{H}(\mathrm{p}, \mathrm{f})$ is a bi-Lyapunov stable then it admits a dominated splitting.

A diffeomorphism f has a heterodimensiional cycle associated with the hyperbolic periodic points $p$ and $q$ of $f$ if (i) the indice of the points $p$ and $q$ are different, and (ii) the stable manifold of $p$ meets the unstable manifold of $q$ and the same holds for the stable manifold of $p$ and the unstable manifold of $q$ (see [6]). We say that $f$ has $C^{1}$ robustly heterodimensional cycle if $f$ has a heterodiemnsional cycle associated with the hyperbolic periodic points $p$ and q of f and there is a $\mathrm{C}^{1}$ neighborhood $\mathcal{U}(\mathrm{f})$ of f such that for any $\mathrm{g} \in \mathcal{U}(\mathrm{f}), \mathrm{g}$ has a heterodimensional cycle associated with the hyperbolic periodic points $p_{g}$ and $q_{g}$, where $p_{g}$ and $q_{g}$ are the continuations of $p$ and $q$ for $g$.

Lemma 2 [6, Corollary 1.15] There is a residual set $\mathcal{T} \subset \operatorname{Diff}(M)$ such that for any $\mathrm{f} \in \mathcal{T}$ and every locally maximal chain recurrence class $\mathrm{C}_{\mathrm{f}}$ of f there are two possibilities: either $\mathrm{C}_{\mathrm{f}}$ is hyperbolic or it has a robustly heterodimensional cycle.

Lemma 3 Let $\mathrm{C}(\mathrm{p}, \mathrm{f})$ admits a dominated splitting $\mathrm{E} \oplus \mathrm{F}$ and let $\mathrm{C}(\mathrm{p}, \mathrm{f})$ be locally maximal. If a chain recurrence class $\mathrm{C}(\mathrm{p}, \mathrm{f})$ has heterodimensional cycle then f does not have the orbital shadowing property on $\mathrm{C}(\mathrm{p}, \mathrm{f})$.

Proof. Suppose, by contradiction, that $f$ has the orbital shadowing property on $C(p, f)$. Since $C(p, f)$ has heterodimensional cycle, there is $q$ a hyperbolic periodic point in $C(p, f)$ such that index $(p) \neq \operatorname{index}(q)$. Then we take $x, y \in$ $C(p, f)$ such that $x \in W^{s}(p) \cap W^{u}(q)$ and $y \in W^{u}(p) \cap W^{s}(q)$. By assumption, $x, y$ are not transverse intersection points. Since $q$ is hyperbolic, $T_{q} M=E_{q}^{s} \oplus$ $\mathrm{E}_{\mathrm{q}}^{\mathrm{u}}$. Choose $\alpha>0$ sufficiently small such that $\mathrm{W}_{\alpha / 4}^{\mathrm{s}}(\mathrm{q})=\exp _{q}\left(\mathrm{E}^{\mathrm{s}}(\alpha / 4)\right)$ and $\boldsymbol{W}_{\alpha / 4}^{u}(\mathbf{q})=\exp _{q}\left(E^{u}(\alpha / 4)\right)$. Then we may assume that $y \in W_{\alpha / 4}^{s}(q)$ and $x \in$ $W_{\alpha / 4}^{u}(q)$. Since $y \in W^{u}(p)$, there is $\eta>0$ such that $y \in B_{\eta}(y) \cap W^{u}(p)$. Take a small arc $\mathcal{J}_{y} \subset B_{\eta}(y) \cap W^{u}(p)$ such that $T_{y} \mathcal{J}_{y}=T_{y} W^{y}(p)$. Since $C(p, f)$ admits a dominated splitting $\mathrm{E} \oplus \mathrm{F}$, we have $\mathrm{T}_{y} \mathcal{J}_{y}=\mathrm{F}_{y}=\mathrm{T}_{y} \mathrm{~W}^{u}(p), \mathrm{F}_{\mathrm{q}} \subset \mathrm{E}_{q}^{u}$ and $\mathrm{E}_{\mathrm{q}}^{s} \subset \mathrm{E}_{\mathrm{q}}$. Put $\mathrm{E}^{\mathrm{u}, 1}=\mathrm{E}_{\mathrm{q}} \oplus \mathrm{E}_{\mathrm{q}}^{\mathrm{u}}$ and $\mathrm{E}^{\mathrm{u}, 2}=\mathrm{F}_{\mathrm{q}}$. Then $\mathrm{E}_{\mathrm{q}}^{\mathrm{u}}=\mathrm{E}_{\mathrm{q}}^{\mathrm{u}, 1} \oplus \mathrm{E}_{\mathrm{q}}^{\mathrm{u}, 2}$, and $W_{\alpha / 4}^{u, 1}(q)=\exp _{q}\left(E_{q}^{u, 1}(\alpha / 4)\right), W_{\alpha / 4}^{u, 2}(q)=\exp _{q}\left(E_{q}^{u, 2}(\alpha / 4)\right)$.

Let $\mathrm{P}^{\mathrm{u}}: \mathrm{B}_{\alpha / 4}(\mathrm{q}) \rightarrow \mathrm{E}_{\mathrm{q}}^{\mathrm{u}}$ and $\mathrm{P}^{s}: \mathrm{B}_{\alpha / 4}(\mathrm{q}) \rightarrow \mathrm{E}_{\mathrm{q}}^{s}$ be the projections parallel to $\mathrm{E}_{\mathrm{q}}^{\mathrm{s}}$ and $\mathrm{E}_{\mathrm{q}}^{\mathrm{u}}$, respectively. Then $\mathrm{P}^{\mathrm{u}}\left(\mathrm{f}^{\mathfrak{n}}\left(\mathcal{J}_{y}\right)\right) \cap \mathrm{B}_{\alpha / 4}(\mathrm{q}) \rightarrow \mathrm{W}_{\alpha / 4}^{u, 1}(\mathrm{q})$ and $P^{s}\left(f^{n}\left(\mathcal{J}_{y}\right)\right) \cap B_{\alpha / 4}(q) \rightarrow q$ as $n \rightarrow \infty$.

Take $\epsilon=\min \left\{\alpha / 4, \eta, d\left(x, W_{\alpha / 4}^{u, 1}(p)\right) / 2\right\}$, and let $0<\delta<\epsilon$ be the number of the orbital shadowing property. Since $y \in W^{s}(q) \cap W^{u}(p)$ and $x \in$ $W^{u}(q) \cap W^{s}(p)$, there are $i_{1}>0$ and $i_{2}>0 \operatorname{such}$ that (i)d( $\left.f^{i_{1}}(y), f^{-i_{1}}(x)\right)<\delta$ and $d\left(f^{-\mathfrak{i}_{2}}(y), f^{\mathfrak{i}_{2}}(x)\right)<\delta$, (ii) $\max \left\{d_{H}\left(P^{s}\left(f^{i_{1}+\mathfrak{j}}\left(\mathcal{J}_{y}\right), q\right), d_{H}\left(P^{u}\left(f^{i_{1}+\mathfrak{j}}\left(\mathcal{J}_{y}\right)\right) \cap\right.\right.\right.$ $\left.\left.B_{\alpha / 4}(q), W_{\alpha / 4}^{u, 1}(q)\right)\right\}<\epsilon$ for all $j \geq 0$, where $d_{H}$ is the Hausdorff metric. Then we have a $\delta$-pseudo orbit as follows:

$$
\begin{aligned}
\xi= & \left\{y, f(y), \ldots, f^{i_{1}-1}(y), f^{-i_{1}}(x), \ldots, f^{-1}(x),\right. \\
& \left.x, f(x), \ldots, f^{i_{2}-1}(x), f^{-i_{2}}(y), \ldots, f^{-1}(y), y\right\} \subset C(p, f) .
\end{aligned}
$$

Since $f$ has the orbital shadowing property on $C(p, f)$ and $C(p, f)$ is locally maximal, there is a point $w \in C(p, f)$ such that

$$
\operatorname{Orb}(w) \subset \mathrm{B}_{\epsilon}(\xi) \text { and } \xi \subset \mathrm{B}_{\epsilon}(\operatorname{Orb}(w)) .
$$

First, we assume that there is $k>0$ such that $f^{k}(w) \in \mathcal{J}_{y} \backslash\{y\}$. Then if $f^{k}(w) \in P^{u}\left(f^{n}\left(\mathcal{J}_{y}\right)\right)$ then since $P^{u}\left(f^{n}\left(\mathcal{J}_{y}\right)\right) \rightarrow W^{u, 1}(q)(n \rightarrow \infty), f^{k+n}(w) \rightarrow$ $W^{u}, 1(q)$ as $n \rightarrow \infty$. Thus there is $j>0$ such that $d\left(f^{k+j}(w), f^{j}(y)\right)>8 \epsilon$ and so, $d\left(f^{k+j}(w), q\right)>2 \epsilon$ which is a contradiction. If $f^{k}(w) \notin P^{u}\left(\mathcal{J}_{y}\right)$ then by $\lambda$-lemma, $f^{n}\left(\mathcal{J}_{y}\right) \rightarrow W^{\mu}(q)$ as $n \rightarrow \infty$. Then there is $l>0$ such that $d\left(f^{k+l}(w), q\right)>4 \epsilon$. Since $x \in W^{u}(q)$, there is $m>0$ such that $d\left(f^{-m}(x), q\right)<$ $\epsilon$ for some $m<i_{2}$. Then we know that $f^{k+l}(w) \notin B_{\epsilon}(\xi)$ which is a contradiction by the orbital shadowing property on $C(p, f)$. Then for all $i \in \mathbb{Z}, f^{i}(w) \notin$ $\mathcal{J}_{\mathrm{y}} \backslash\{\mathrm{y}\}$.

We assume that there is $k>0$ such that $f^{k}(w)=y$. Since $y \in W^{s}(q) \cap W^{u}(p)$ and $x \in \mathcal{W}^{u}(q) \cap W^{s}(p)$, we know $\operatorname{Orb}(x) \cap \operatorname{Orb}(y)=\emptyset$. Then we have $\xi \not \subset \mathrm{B}_{\epsilon}(\operatorname{Orb}(w))$ which is a contradiction by the orbital shadowing property on $C(p, f)$. Thus we know $\operatorname{Orb}(w) \cap \mathcal{J}_{y}=\emptyset$.

Finally, we assume that there is $k>0$ such that $f^{k}(w) \in B_{\eta}(y) \backslash \mathcal{J}_{y}$. Then for all $z \in B_{\eta}(y) \backslash \mathcal{J}_{y}$, there is $k>0$ such that $d\left(f^{-k}(x), f^{k}(z)\right)>2 \epsilon$ since $x \in W^{\mu}(q)$ and $q$ is hyperbolic saddle. Then we have $\xi \not \subset B_{\epsilon}(\operatorname{Orb}(w))$ which is a contradiction by the orbital shadowing property on $C(p, f)$. Consequently, if a locally maximal chain recurrence class $C(p, f)$ admits a dominated splitting and $f$ has the orbital shadowing property on $C(p, f)$ then it does not the heterodimensional cycle.

Proof of Theorem A. Let $\mathrm{f} \in \mathcal{G} \cap \mathcal{T}$ and let f has the orbital shadowing property on a locally maximal chain transitive set $\mathcal{C}$. Since $f \in \mathcal{G}$, by Remark $1 \mathcal{C}=C(p, f)$ for some hyperbolic periodic point $p$. Since chain transitive set $\mathcal{C}$ admits a dominated splitting $E \oplus F$, and $f$ has the orbital shadowing property
on a locally maximal chain transitive $\mathcal{C}$, by Lemma $3, \mathcal{C}=\mathrm{C}(\mathrm{p}, \mathrm{f})$ does not have the heterodimensional cycles. It is clear that a locally maximal $C(p, f)$ does not have the robustly hetrodimensional cycle. Thus by Lemma 2, a locally maximal chain transitive set $\mathcal{C}$ is hyperbolic.

In Abdenur et al [2] the authors proved that every locally maximal homoclinic class with a non-empty interior is the whole space.

Corollary 1 Let $\mathrm{f}: \mathrm{M} \rightarrow \mathrm{M}$ be a diffeomorphism with $\operatorname{dim} M=3$. For $\mathrm{C}^{1}$ generic f , if f has the orbital shadowing property on a locally maximal chain transitive set $\mathcal{C}$ which admits a dominated splitting $\mathrm{E} \oplus \mathrm{F}$ then it is Anosov.

The following was proved in [21, Proposition 1.2] which means that a homoclinic class admits a codimension one dominated splitting then it has a non-empty interior.

Lemma 4 There is a residual set $\mathcal{R} \subset \operatorname{Diff}\left(M^{3}\right)$ such that for any $\mathrm{f} \in \mathcal{R}$, if a homoclinic class H admits a codimension one dominated splitting then it has non-empty interior.

Proof of Corollary 1. Let $\mathrm{f} \in \mathcal{G} \cap \mathcal{T} \cap \mathcal{R}$ and let f has the orbital shadowing property on a locally maximal chain transitive set $\mathcal{C}$ which admits a dominated splitting $E \oplus F$. By Remark 1, a locally maximal chain transitive $\mathcal{C}=H(p, f)$. Since $\mathrm{f} \in \mathcal{R}$, by Lemma 4, a hoomoclinic class $\mathrm{H}(\mathrm{p}, \mathrm{f})$ has nonempty interior. Since $H(p, f)$ is locally maximal, by $[2$, Thereom 3], $H(p, f)=M$. Since $f$ has the orbital shadowing property on a locally maximal chain transitive set $\mathcal{C}$ which admits a dominated splitting $\mathrm{E} \oplus \mathrm{F}$, by Theorem A , it is hyperbolic, and so it is Anosov.

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## References

[1] F. Abdenur, C. Bonatti and S. Crovisier, Nonuniform hyperbolicty for $C^{1}$-generic diffeomorphisms, Isral J. Math., 183 (2011), 1-60.
[2] F. Abdenur, C. Bonatti and L. J. Díaz, Non-wandering sets with non empty interior, Noinearity, 17 (2004), 175-191.
[3] F. Abdenur and L. J. Díaz, Pseudo-orbit shadowing in the $\mathrm{C}^{1}$-topology, Discrete Contin. Dyn. Syst., 17 (2) (2007), 223-245.
[4] J. Ahn, K. Lee and M. Lee, Homoclinic classes with shadowing, J. Inequal. Appl., 2012:97 (2012), 1-6.
[5] C. Bonatti and S. Crovisier, Récurrence et généricité, Invent. Math., 158 (2004), 33-104.
[6] C. Bonatti and L. J. Díaz, Robust heterodimensional cycles and $C^{1}$ generic dynamics, J. Inst. Math. Jessieu, 7 (2008), 469-525.
[7] C. Carballo, C. A. Morales and M. J. Pcaifico, Homoclinic classes for C ${ }^{1}$ generic vector fields, Ergod. Theorey Dynam. Syst., 23 (2003), 403-415.
[8] S. Crovisier, Periodic orbits and chain transitive sets of $\mathrm{C}^{1}$-diffeomorphisms, Publ. Math. Inst. Hautes Etudes. Sci., 104 (2006), 87-141.
[9] K. Lee and M. Lee, Shadowable chain recurrence classes for generic diffeomorphisms, Taiwan J. Math., 20 (2016), 399-409.
[10] K. Lee and M. Lee, Volume preserving diffeomorphisms with orbital shadowing, J. Inequal. Appl., 2013:18, (2013), 5 pp.
[11] K. Lee and X. Wen, Shadowable chain transitive sets of $C^{1}$ generic diffeomorphisms, Bull. Korean Math. Soc. 49 (2012), 263-270.
[12] M. Lee, Hamiltonian systems with orbital, orbital inverse shadowing, Adv. Difference Equ., 2014:192(2014), 9 pp.
[13] M. Lee, Orbital shadowing property for generic divergence-free vector fields, Chaos Solitons \& Fractals, 54 (2013), 71-75.
[14] M. Lee, Orbital shadowing for $\mathrm{C}^{1}$-generic volume-preserving diffeomorphisms, Abstr. Appl. Anal., 2013, Art. ID 693032, 4 pp.
[15] M. Lee, Divergence-free vector fields with orbital shadowing, Adv. Difference Equ, 2013:132, (2013), 6 pp.
[16] M. Lee, Robustly chain transitive sets with orbital shadowing diffeomorphisms, Dyn. Syst., 27 (2012), 507-514.
[17] M. Lee, Chain components with $\mathrm{C}^{1}$-stably orbital shadowing, Adv. Difference Equ., 2013:67 (2013), 12 pp.
[18] M. Lee, Chain transitive sets with dominated splitting, J. Math. Sci. Adv. Appl., 4 (2010), 201-208.
[19] A. V. Osipov, Nondensity of the orbital shadowing property in $\mathrm{C}^{1}$ topology, (Russian) Algebra i Analiz 22 (2010), no. 2, 127-163; translation in St. Petersburg Math. J., 22 (2) (2011), 267-292.
[20] S. Y. Pilyugin, A. A. Rodionova and K. Sakai, Orbital and weak shadowing properties, Discrete Contin. Dyn. Syst., 9 (2) (2003), 287-308.
[21] R. Potrie, Generic bi-Lyapunov stable homoclinic classes, Nonlinearity, 23 (2010), 1631-1649.
[22] C. Robinson, Stability theorem and hyperbolicity in dynamical systems, Rocky Mountain J. Math., 7 (1977), 425-437.
[23] K. Sakai, Sahdowable chain transitive sets, J. Diff. Eqaut. Appl., 19 (2013), 1601-1618.
[24] K. Sakai, Pseudo orbit tracing property and strong transversality of diffeomorphisms on closed manifolds, Osaka J. Math., 31 (1994), 373-386.

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# Restrained domination in signed graphs 

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#### Abstract

A signed graph $\Sigma$ is a graph with positive or negative signs attatched to each of its edges. A signed graph $\Sigma$ is balanced if each of its cycles has an even number of negative edges. Restrained dominating set D in $\Sigma$ is a restrained dominating set of its underlying graph where the subgraph induced by the edges across $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}]$ and within $\mathrm{V} \backslash \mathrm{D}$ is balanced. The set D having least cardinality is called minimum restrained dominating set and its cardinality is the restrained domination number of $\Sigma$ denoted by $\gamma_{\mathrm{r}}(\Sigma)$. The ability to communicate rapidly within the network is an important application of domination in social networks. The main aim of this paper is to initiate a study on restrained domination in the realm of different classes of signed graphs.


[^6]
## 1 Introduction

Graphs used in this article, unless otherwise mentioned will be undirected, simple and finite. For a graph $G=(V, E)$, the degree of a vertex $v$, denoted by $\operatorname{deg}(v)$ is the number of edges incident to the vertex $v$ (loops counted twice incase of multigraph). The maximum degree of G is denoted by $\Delta(\mathrm{G})$ and the minimum degree of G is denoted by $\delta(\mathrm{G})$. If $\operatorname{deg}(v)=1$, then $v$ is called a pendant vertex. For all terminology and notation in graph theory, we refer the reader to the text-book by Harary [1]. The graphs with positive or negative signs attatched to each of its arcs are called signed graphs. Zaslavasky [2], formally defines a signed graph as $\Sigma=(G, \sigma)$, where $G$ is the underlying unsigned graph consisting of $G=(\mathrm{V}, \mathrm{E})$ and $\sigma: \mathrm{E} \rightarrow\{+,-\}$ is the function assigning signs to the edges of the graph. The edges which receive $+(-)$ signs, are called positive(negative) edges of $\Sigma$.

A signed graph $\Sigma$ is all-positive(all-negative) if all its edges are positive (negative). If it is an all-positive or all-negative, then it is said to be homogenous else heterogenous. Switching $\Sigma$ with respect to a marking $\mu$ where $\mu: \mathrm{V} \rightarrow\{+1,-1\}$ is the operation of negating every edge whose end vertices are of opposite signs. $\Sigma$ is said to be balanced if each of its cycle has an even number of negative edges. Equivalently, a signed graph is balanced if it can be switched to an all-positive signed graph. For further details on theory of signed graphs, the reader is referred to [3, 2].

Domination in graph theory for unsigned graphs is one of the continuing research of the well-researched region. Detailed survey of the same can be found in the book by Haynes et al. [4]. In 2013, Acharya [5] introduced the theory of dominance for signed graphs as well as signed digraphs. A subset $\mathrm{D} \subseteq \mathrm{V}$ of vertices of $\Sigma=(\mathrm{G}, \sigma)$ is a dominating set of $\Sigma$, if there exists a marking $\mu: V \rightarrow\{+1,-1\}$ of $\Sigma$ such that every vertex not in D is adjacent to at least one vertex in $D$ and $\sigma(u v)=\mu(u) \mu(v), \forall u \in V \backslash D$. The minimum cardinality of a dominating set in $\Sigma$ is called its domination number, denoted by $\gamma(\Sigma)$. Germina and Ashraf $[6,7]$ gave characterization for open domination and double domination in signed graphs. In 2015, Walikar et al. [8] introduced the concept of signed domination for signed graphs.

In a social network, if all individuals are connected to at least one such person who can be reached directly, an emergency message can easily be sent to all participants in the network, thus reducing delay time. Nevertheless, it is also important to examine positive and negative relationships between individuals when examining social network interactions. This situation can be modeled on what is known as the dominating set problem in signed graphs.

In this paper, we introduce the concept of restrained domination for signed graphs. In addition, we determine the best possible bounds on $\gamma_{r}(G)$ for certain classes of signed graphs.

## 2 Definitions and results

The concept of restrained domination in graphs was introduced by Domke et al. [9] in 1999. A set $\mathrm{D} \subseteq \mathrm{V}$ is a restrained dominating set of graph $\mathrm{G}=(\mathrm{V}, \mathrm{E})$, if every vertex in $V \backslash \mathrm{D}$ is adjacent to a vertex in D as well as another vertex in $\mathrm{V} \backslash \mathrm{D}$. The restrained domination number of graph G denoted by $\gamma_{\mathrm{r}}(\mathrm{G})$ is the smallest cardinality of a restrained dominating set of G . We will now define the concept of restrained domination for signed graphs and then find the best possible general bounds for some classes of signed graph. In this paper, we will be using the notation $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}]$ when $\mathrm{D} \subseteq \mathrm{V}$, to denote the subgraph of $\Sigma$ induced by the edges of $\Sigma$ with one end point in D and the other end point in $V \backslash \mathrm{D}$. Induced subgraph in $\mathrm{V} \backslash \mathrm{D}$ is denoted by $\Sigma[\mathrm{V} \backslash \mathrm{D}]$.

Definition $1 A$ subset $\mathrm{D} \subseteq \mathrm{V}$ of vertices of $\Sigma=(\mathrm{V}, \mathrm{E}, \sigma)$ is a restrained dominating set if there exists a marking $\mu: \mathrm{V} \rightarrow\{+1,-1\}$ of $\Sigma$ such that every vertex in $\mathrm{V} \backslash \mathrm{D}$ is adjacent to a vertex in D as well a vertex in $\mathrm{V} \backslash \mathrm{D}$ and for every vertex $u$ in $\mathrm{V} \backslash \mathrm{D}, \sigma(u v)=\mu(u) \mu(v) \forall v \in \mathrm{D}$ and $v \in \mathrm{~V} \backslash \mathrm{D}$.

The minimum cardinality of a restrained dominating set D is called restrained domination number of $\Sigma$ denoted by $\gamma_{\mathrm{r}}(\Sigma)$. Every restrained dominating set of a signed graph $\Sigma$ of order $n$ follows the inequality $1 \leq \gamma_{r}(\Sigma) \leq n$. As each vertex in $\mathrm{V} \backslash \mathrm{D}$ is adjacent to at least one other vertex in $\mathrm{V} \backslash \mathrm{D}$, the cardinality of the set $\mathrm{V} \backslash \mathrm{D}$ is always greater than or equal to two. Hence, $\gamma_{r}(\Sigma)$ can never be equal to $n-1$. Before proceeding further with results on bounds for few classes of signed graphs, we state some important results used for obtaining these bounds.

Proposition 1 Let G be any graph of order n with $\delta(\mathrm{G})=1$. Then, every restrained dominating set of graph G must necessarily have all its pendant vertices.

Switching $\Sigma=(G, \sigma)$ with respect to $\mu$ means forming the switched graph $\Sigma^{\mu}=\left(\Sigma, \sigma^{\mu}\right)$, whose underlying graph is the same but whose sign function is defined on an edge $e: \nu w$ by $\Sigma^{\mu}(e)=\mu(v) \sigma(e) \mu(w)$. In case of balanced signed graphs, when $\Sigma$ is switched with respect to $\mu$, we obtain an all-positive signed graph. Hence, we can state the following lemma:

Lemma 1 [2] A signed graph $\Sigma$ is balanced if and only if it can be switched to an all-positive signed graph.

Let, $D_{\Sigma}^{r}$ be the set of all restrained dominating sets of signed graphs and $D_{|\Sigma|}^{r}$ be the set of all restrained dominating sets of its underlying graphs. Then, for balanced signed graphs $\Sigma, \mathrm{D}_{\Sigma}^{r}=\mathrm{D}_{|\Sigma|}^{\mathrm{r}}$. But, note that this equality does not hold true for all unbalanced signed graphs. For example, consider a 6-cycle graph $\Sigma=C_{6}$ with three negative edges, denoted as $C_{6}^{(3)}$. The underlying graph $|\Sigma|=$ $\mathrm{C}_{6}$ will have all independent vertices in its minimum restrained dominating set. Whereas, in $\Sigma$ there exists no such independent vertices in restrained dominating set. This leads to the conclusion of following proposition:

Proposition 2 The set $\mathrm{D}_{\Sigma}^{r}$ of all restrained dominating sets of a signed graph $\Sigma$ is contained in the set $\mathrm{D}_{|\Sigma|}^{\mathrm{r}}$ of all restrained dominating sets of its underlying graph $|\Sigma|$.

Proposition 3 For any finite balanced signed graph, $\gamma_{\mathrm{r}}(\Sigma)=\gamma_{\mathrm{r}}(|\Sigma|)$.
Clearly, from Lemma 1, we can conclude for balanced signed graphs $\mathrm{D}_{\Sigma}^{r}=$ $D_{|\Sigma|}^{r}$ and hence the result holds true.

Further, since $D_{\Sigma}^{\mathrm{r}} \subseteq \mathrm{D}_{|\Sigma|}^{\mathrm{r}}$, we can conclude $\gamma_{\mathrm{r}}(|\Sigma|) \leq \gamma_{\mathrm{r}}(\Sigma)$. In the following results, we derive bounds for some classes of signed graphs.

Theorem 1 If $\mathrm{P}_{n}^{(\mathrm{r})}$ is a signed path with $n$ vertices and $r$ negative edges, $\gamma_{\mathrm{r}}\left(\mathrm{P}_{\mathrm{n}}^{(\mathrm{r})}\right)=\gamma_{\mathrm{r}}\left(\mathrm{P}_{\mathrm{n}}\right)=\mathrm{n}-2\lfloor(\mathrm{n}-1) / 3\rfloor$ for $\mathrm{n} \geq 1$ and $0<\mathrm{r} \leq \mathrm{n}-1$.

Proof. In the case of restrained dominating set of $P_{n}$, it is proved in [9] that $\gamma_{r}\left(P_{n}\right)=n-2\lfloor(n-1) / 3\rfloor$. Since signed paths are trivially balanced, then by Proposition 3, the theorem follows.

Theorem 2 If $\mathrm{C}_{\mathrm{n}}^{(\mathrm{r})}$ is a signed cycle with $\mathrm{n} \equiv 1,2(\bmod 3)$ and $0<\mathrm{r} \leq \mathrm{n}$, then $\gamma_{\mathrm{r}}\left(C_{n}^{(r)}\right)=\gamma_{r}\left(C_{n}\right)=n-2\lfloor n / 3\rfloor$.

Proof. Let $\Sigma$ be a signed cycle $C_{n}^{(r)}$ with $n \equiv 1,2(\bmod 3)$ and $0<r \leq n$. Restrained domination number as proved in [9] for $C_{n}$ is $n-2\lfloor n / 3\rfloor$. We consider following two different cases to derive bounds for $\gamma_{r}\left(C_{n}\right)$ :

Case $1 r \equiv 0(\bmod 2)$.

Since, $\Sigma$ has an even number of negative edges, therefore $\Sigma$ is balanced. Thus by Proposition 3, $\gamma_{r}\left(C_{n}\right)=\gamma_{r}(\Sigma)=n-2\lfloor n / 3\rfloor$.

Case $2 r \equiv 1(\bmod 2)$.
Let D be the minimum restrained dominating set of the underlying graph $|\Sigma|$. We need to check if D is a minimum restrained dominating set of $\Sigma$ also. Let us suppose, D is restrained dominating set of $\Sigma$. Then, there exists at least one pair of adjacent vertices in D . Thus, $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}] \cup \Sigma[\mathrm{V} \backslash \mathrm{D}]$ will always be acyclic, which is trivially balanced. Since, set D satisfies the property given in the definition, therefore it is a restrained dominating set of $\Sigma$. We know, $\mathrm{D}_{\Sigma}^{r} \subseteq \mathrm{D}_{[\Sigma \mid}^{r}$ by Proposition 2. We can thus conclude, D is minimum restrained dominating set of $\Sigma$ and hence follows the theorem.

Theorem 3 Let $\mathrm{C}_{\mathrm{n}}^{(\mathrm{r})}$ be a signed cycle with $\mathrm{n} \equiv 0(\bmod 3)$ and $0<r \leq n$, then

$$
\gamma_{r}\left(C_{n}^{(r)}\right)= \begin{cases}n-2\lfloor n / 3\rfloor & \text { if } r \text { is even } \\ n-2\lfloor n / 3\rfloor+2 & \text { if } r \text { is odd }\end{cases}
$$

Proof. Let $\Sigma$ be a signed cycle $C_{n}^{(r)}$ with $n \equiv 0(\bmod 3)$ and $0<r \leq n$. Proceeding in a similar way as previous theorem, we form two cases for $\gamma_{r}\left(C_{n}\right)$ :

Case $1 r \equiv O(\bmod 2)$
Since, $\Sigma$ has an even number of negative edges, therefore $\Sigma$ is balanced. Thus by Proposition 3, $\gamma_{r}\left(C_{n}\right)=\gamma_{r}(\Sigma)=n-2\lfloor n / 3\rfloor$.
Case $2 r \equiv 1$ (mod 2)
Let D be the minimum restrained dominating set of the underlying graph $|\Sigma|$. We need to check, if D is restrained dominating set of $\Sigma$. In this case we observe that the set D has all vertices at a distance three from each other. Therefore, $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}] \cup \Sigma[\mathrm{V} \backslash \mathrm{D}]$ will be a cycle with odd number of negative edges and hence not balanced. Thus by definition, D will not be a restrained dominating set of $\Sigma$. We now need to add more vertices to $D$. Suppose, a vertex $v_{1} \in V \backslash D$ is added to the set D . The neighboring vertex of $v_{1}$ in $\mathrm{V} \backslash \mathrm{D}$, say $v_{2}$ then has no neighboring vertex in $\mathrm{V} \backslash \mathrm{D}$ and is not a restrained dominating set. Thus, we will need to add more vertices to D . Let us add $\mathrm{N}\left(v_{1}\right) \in \mathrm{V} \backslash \mathrm{D}$ to the set D. Then, there exists only signed paths in $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}] \cup \Sigma[\mathrm{V} \backslash \mathrm{D}]$, which is trivially balanced. Since, we added two more vertices to the set D, therefore $\gamma_{r}(\Sigma)=\gamma_{r}\left(C_{n}\right)+2=n-2\lfloor n / 3\rfloor+2$.

Theorem 4 If $\mathrm{K}_{1, n-1}^{(\mathrm{r})}$ is a star signed graph with $n$ vertices and $r$ negative edges, then $\gamma_{\mathrm{r}}\left(\mathrm{K}_{1, \mathrm{n}-1}^{(\mathrm{r})}\right)=\mathrm{n}$.

Proof. Let $\Sigma$ be a star signed graph $K_{1, n-1}^{(r)}$ with $r$ negative edges. Since, $\gamma_{\mathrm{r}}\left(\mathrm{K}_{1, \mathrm{n}-1}\right) \leq \gamma_{\mathrm{r}}\left(\mathrm{K}_{1, n-1}^{(\mathrm{r})}\right) \leq \mathrm{n}$ and $\gamma_{\mathrm{r}}\left(\mathrm{K}_{1, n-1}\right)=\mathrm{n}$, the theorem holds true.

For complete signed graph $\mathrm{K}_{\mathrm{n}}, \mathrm{n} \geq 5$, we can derive a general bounds as shown in Theorem 5 to obtain $\gamma_{r}$. But, $\mathrm{K}_{4}^{(\mathrm{r})}$ doesnot satisfy this theorem. Hence, we state the following proposition.

Note that, paw graph is the graph obtained by joining a vertex of cycle graph $C_{3}$ to a singleton graph $K_{1}$. In the following proposition, $P_{2} \cup P_{2}$ is the union of two disconnected paths $\mathrm{P}_{2}$.

Proposition 4 Let $\Sigma$ be a $\mathrm{K}_{4}^{(\mathrm{r})}$ graph with $r$ negative edges and $r$ is even and let $\langle\mathrm{I}\rangle$ be all-negative edge induced subgraph of $\Sigma$. Then,

$$
\gamma_{\mathrm{r}}(\Sigma)= \begin{cases}1, & \text { if }\langle\mathrm{I}\rangle \cong \mathrm{C}_{4} \\ 2, & \text { if }\langle\mathrm{I}\rangle \cong \mathrm{P}_{3} \text { or }\langle\mathrm{I}\rangle \text { is a paw graph } \\ 4, & \text { if }\langle\mathrm{I}\rangle \cong \mathrm{K}_{4}^{(6)} \text { or }\langle\mathrm{I}\rangle \cong \mathrm{P}_{2} \cup \mathrm{P}_{2} .\end{cases}
$$

Theorem 5 If $p$ is the order of the subgraph induced by negative edges of a complete signed graph $\mathrm{K}_{\mathrm{n}}$ with $n$ vertices, $n \geq 5$, then

$$
\gamma_{r}\left(K_{n}^{(r)}\right)= \begin{cases}p & \text { if } p<n-1 \\ n & \text { otherwise }\end{cases}
$$

Proof. Let $\Sigma$ be any complete signed graph having $r$ negative edges and $n \geq 5$ and D be the minimum restrained dominating set of $\Sigma$. We need to show that all the vertices incident to any negative edge in $\Sigma$ belongs to the set D . We prove this by contradiction. Suppose there exists at least one negative edge in $\Sigma$ with end vertices say $\nu_{1}$ and $\nu_{2}$, such that either both or one end vertex is not in D . Then, the negative edge $\nu_{1} \nu_{2}$ will either be in $\Sigma[\mathrm{V} \backslash \mathrm{D}]$ or $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}]$. Now, there exists at least one $\mathrm{C}_{3}$ in $\Sigma[\mathrm{V} \backslash \mathrm{D}]$ or $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}] \cup \Sigma[\mathrm{V} \backslash \mathrm{D}]$ having odd number of negative edges, and thus $\Sigma$ is not balanced. This implies, by Definition 1 that $D$ is not a minimum restrained dominating set. $D$ satisfies Definition 1 only when there does not exists any negative edge in $\Sigma[\mathrm{V} \backslash \mathrm{D}]$ or $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}]$, which contradicts our assumption. Therefore, $\gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathrm{n}}^{(\mathrm{r})}\right)=\mathrm{p}$, for $\mathrm{p}<\mathrm{n}-1$. By Definition $1, \gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathrm{n}}^{(\mathrm{r})}\right)$ can never be equal to $\mathrm{n}-1$.

Hence, $\gamma_{r}\left(K_{n}^{(r)}\right)>n-1$ for $p>n-1$. But, we know $\gamma_{r}\left(K_{n}^{(r)}\right) \leq n$. Therefore, $\gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathrm{n}}^{(\mathrm{r})}\right)=\mathrm{n}$.

Restrained domination for complete bipartite signed graph $\Sigma$ varies based on the number of negative edges in $\Sigma$ and hence, for large graphs it is difficult to find exact restrained domination number. In the following theorems, we have generalized some of those cases for complete bipartite signed graphs $K_{m, m}^{(r)}$ with $2 m$ vertices and concluded by giving the bounds on $\gamma_{r}$ for any complete bipartite signed graph.

Theorem 6 Let $\mathrm{K}_{\mathrm{m}, \mathrm{m}}^{(\mathrm{r})}$ be a complete bipartite signed graph with 2 m vertices and r negative edges and $\langle\mathrm{I}\rangle$ denote the subgraph induced by all negative edges, then $\gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathrm{m}, \mathrm{m}}^{(\mathrm{r})}\right)=2$ in any one of the following conditions:

1. $\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}-1}$ or $\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}}$.
2. All the edges are negative, i.e. $\langle\mathrm{I}\rangle \cong \mathrm{K}_{\mathrm{m}, \mathrm{m}}^{(\mathrm{r})}$, where $\mathrm{r}=\mathrm{m}$

Proof. Let $\Sigma$ be a complete bipartite graph $K_{m, m}^{(r)}$ with $2 m$ vertices and $r$ negative edges and D be the restrained dominating set of $\Sigma$. We denote $\langle\mathrm{I}\rangle$ for the subgraph induced by all negative edges of $\Sigma$.

Case $1\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}-1}$ or $\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}}$
Any induced cycle in a complete bipartite graph is always even. Also, for a signed graph to be balanced, every cycle in the graph must have an even number of negative edges. Moreover, degree of every vertex in cycle is always 2. Let $u$ be the vertex to which all the negative edges are incident. All the induced cycles of $\Sigma$ not including vertex $u$ are all positive and hence satisfy the marking $\sigma(v w)=\mu(v) \mu(w) \forall v, w \neq u$. Thus, we need to check for the induced cycles in $\Sigma$ containing the vertex $\mathfrak{u}$. In case of $\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}}, \Sigma$ can be switched to all positive signed graph, and hence, by Proposition 3, $\gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathrm{m}, \mathrm{m}}\right)=\gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathrm{m}, \mathrm{m}}^{(\mathrm{r})}\right)=2$ for $\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}}$. In case of $\langle\mathrm{I}\rangle \cong \mathrm{K}_{1, \mathrm{~m}-1}$, every cycle incuding vertex $u$ will either have two negative edges incident to vertex $\boldsymbol{u}$, which is always an even cycle or it will have 1 negative and 1 positive edge incident to $u$, which is an odd cycle. Hence, in case of odd cycle if we take the end vertices of the positive edge incident to vertex $u$ in the set D , we get the desired result.

Case $2\langle\mathrm{I}\rangle \cong \mathrm{K}_{\mathrm{m}, \mathrm{m}}$.

This implies $\Sigma$ is all negative and switching graph $\Sigma$, we obtain all positive $\mathrm{K}_{\mathrm{m}, \mathrm{m}}$ graph. Thus, by Lemma $1, \gamma_{\mathrm{r}}(\Sigma)=2$.

Theorem 7 Let $\Sigma$ be a complete bipartite signed graph $\mathrm{K}_{\mathrm{m}, \mathrm{m}}^{(\mathrm{r})}$ with 2 m vertices and r negative independent edges with $\mathrm{r} \leq \mathrm{m}$. Then

$$
\gamma_{r}(\Sigma)= \begin{cases}2 r, & \text { if } r<m \\ 2(r-1), & \text { if } r=m\end{cases}
$$

Proof. Let $\Sigma$ be $K_{m, m}^{(r)}$ with $2 m$ vertices and $r$ negative independent edges with $r \leq m$. Let $D$ be the restrained dominating set of $\Sigma$. Since, all the negative edges in $\Sigma$ are independent, therefore no two negative edges have at least one end point in common. In this case, there always exists at least one cycle in $\Sigma$ containing odd number of negative edges and hence $\Sigma$ is not balanced. Thus, to satisfy Definition 1 we will need to choose all the end points of the negative edges in the set D such that $\Sigma[\mathrm{D}: \mathrm{V} \backslash \mathrm{D}] \cup \Sigma[\mathrm{V} \backslash \mathrm{D}]$ is balanced.

Case 1 Suppose, there are $r$ independent negative edges with $r<m$. Then, number of vertices in D will be twice the number of negative edges and hence $\gamma_{r}(\Sigma)=2 r$ for $r<m$.

Case 2 Now, suppose that the number of independent negative edges $r$ is equal to $m$. Then, $D$ must include all the vertices of $\Sigma$ and hence, $\gamma_{r}(\Sigma)$ must be equal to $2 r$. But, this is not the minimum restrained domination number and hence, we need to remove some vertices from the set D. Since, $\gamma_{r}$ cannot be equal to $2 m-1$, we will remove two vertices from set $D$. The set $D$ is now minimum restrained dominating set. Thus, $\gamma_{\mathrm{r}}(\Sigma)=2(\mathrm{r}-1)$.

Thus we can conclude with the following corollary:
Corollary 1 Let $\Sigma$ be any complete bipartite signed graph $\mathrm{K}_{\mathrm{m}, \mathrm{m}}^{(\mathrm{r})}$ with 2 m vertices and r negative edges, then $2 \leq \gamma_{\mathrm{r}}\left(\mathrm{K}_{\mathfrak{m}, \mathfrak{m}}^{(\mathrm{r})}\right) \leq 2(\mathrm{~m}-1)$.

## 3 Conclusion

In this paper, we introduced the concept of restrained domination for signed graphs and determined the bounds on $\gamma_{\mathrm{r}}(\Sigma)$ for certain classes of signed graphs. As for further work, it can be extended in finding bounds for other classes of derived signed graphs. Also, it would be interesting to study on
the critical concepts of restrained domination in signed graphs. The above concepts is very much useful in fault tolerence analysis of communication networks, social networks and security systems.

## References

[1] F. Harary, Graph theory, Addison-Wesley, Reading, MA, 1969.
[2] T. Zaslavsky, Signed graphs, Discrete Applied Mathematics, 4 (1982), 47-74.
[3] F. Harary, On the notion of balance of a signed graph, The Michigan Mathematical Journal, 2 (1953), 143-146.
[4] T. W. Haynes, S. Hedetniemi, P. Slater, Fundamentals of domination in graphs, CRC press, 2013.
[5] B. D. Acharya, Domination and absorbance in signed graphs and digraphs. I: Foundations, The Journal of Combinatorial Mathematics and Combinatorial Computing, 84 (2013), 1-10.
[6] K. A. Germina, P. Ashraf, On open domination and domination in signed graphs, International Mathematical Forum, 8 (2013), 1863-1872.
[7] K. A. Germina, P. Ashraf, Double domination in signed graphs, Cogent Mathematics, 3 (2016), 1-9.
[8] H. B. Walikar, S. Motammanavar, B. D. Acharya, Signed domination in signed graphs, Journal of Combinatorics and System Sciences, 40 (2015), 107-128.
[9] G. S. Domke, J. H. Hattingh, S. T. Hedetniemi, R. C. Laskar, L. R. Markus, Restrained domination in graphs, Discrete Mathematics, 203 (1999), 61-69.

# Graded Morita theory over a G-graded G-acted algebra 

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#### Abstract

We develop a group graded Morita theory over a G-graded G-acted algebra, where $G$ is a finite group.


## 1 Introduction

Let $G$ be a finite group. In this article we develop a group graded Morita theory over a G-graded G-acted algebra, which is motivated by the problem to give a group graded Morita equivalences version of relations between character triples (see [11]) as in [9]. We will follow, in the development of graded Morita theory over a G-graded G-acted algebra, the treatment of Morita theory given by C. Faith in 1973 in [5]. Significant in this article is the already developed graded Morita theory. Graded Morita theory started in 1980 when R. Gordon and E. L. Green have characterized graded equivalences in the case of $G=\mathbb{Z}$, in [6]. Furthermore, in 1988 it was observed to work for arbitrary groups G by C. Menini and C. Năstăsescu, in [10]. We will make use of their results under the form given by A. del Río in 1991 in [4] and we will also use the graded Morita theory developed by P. Boisen in 1994 in [2].

[^7]This paper is organized as follows: In Section $\S 2$, we introduce the general notations. In Section §3, we recall from [9] the definition of a G-graded Gacted algebra, we fix one to which we will further refer to by $\mathscr{C}$, we recall the definition of a G-graded algebra over said $\mathscr{C}$, and we recall the definition of a graded bimodule over $\mathscr{C}$. Moreover, we will give some new examples, useful for this article, for each notion. In Section $\S 4$, we construct the notion of a G-graded Morita context over $\mathscr{C}$ and we will give an appropriate example. In Section $\S 5$, we introduce the notions of graded functors over $\mathscr{C}$ and of graded Morita equivalences over $\mathscr{C}$ and finally we state and prove two Morita-type theorems using the said notions.

## 2 Notations and preliminaries

Throughout this article, we will consider a finite group G. We shall denote its identity by 1 .

All rings in this paper are associative with identity $1 \neq 0$ and all modules are unital and finitely generated. We consider $\mathcal{O}$ to be a commutative ring.

Let $\mathcal{A}$ be a ring. We denote by $\mathcal{A}$-Mod the category of all left $\mathcal{A}$-modules. We shall usually write actions on the left, so in particular, by module we will usually mean a left module, unless otherwise stated. The notation ${ }_{A} M$ (respectively, ${ }_{A} M_{A^{\prime}}$ ) will be used to emphasize that $M$ is a left $A$-module (respectively, an ( $A, A^{\prime}$ )-bimodule).
Let $A=\bigoplus_{g \in G} A_{g}$ be a $G$-graded $\mathcal{O}$-algebra. We denote by $A$ - Gr the category of all G -graded left A -modules. The forgetful functor from $\mathrm{A}-\mathrm{Gr}$ to A -Mod will be denoted by $U$. For $M=\bigoplus_{g \in G} M_{g} \in A$ - $G r$ and $g \in G$, the $g$-suspension of $M$ is defined to be the $G$-graded $A$-module $M(g)=\oplus_{h \in G} M(g)_{h}$, where $M(g)_{h}=M_{g h}$. For any $g \in G, T_{g}^{A}: A-G r \rightarrow A-G r$ will denote (as in [4]) the g -suspension functor, i.e. $\mathrm{T}_{\mathrm{g}}^{\mathcal{A}}(\mathrm{M})=M(\mathrm{~g})$ for all $\mathrm{g} \in \mathrm{G}$. The stabilizer of $M$ in G is, by definition $[7, \S 2.2 .1]$, the subgroup

$$
\mathrm{G}_{\mathrm{M}}=\{\mathrm{g} \in \mathrm{G} \mid \mathrm{M} \simeq \mathrm{M}(\mathrm{~g}) \text { as } G \text {-graded left } A \text {-modules }\} .
$$

Let $M, N \in A$-Gr. We denote by $\operatorname{Hom}_{\mathcal{A}}(M, N)$, the additive group of all $A$ linear homomorphisms from $M$ to $N$. Because $G$ is finite, E. C. Dade showed in [3, Corollary 3.10] that $\operatorname{Hom}_{\mathcal{A}}(\mathrm{M}, \mathrm{N})$ is G -graded. More precisely, if $\mathrm{g} \in \mathrm{G}$, the component of degree g (furthermore called the g -component) is defined as in $[7,1.2]$ :

$$
\operatorname{Hom}_{\mathcal{A}}(M, N)_{g}:=\left\{f \in \operatorname{Hom}_{\mathcal{A}}(M, N) \mid f\left(M_{x}\right) \subseteq N_{x g}, \text { for all } x \in G\right\} .
$$

We denote by $i^{\prime} X$ the identity map defined on a set $X$.

## 3 Graded bimodules over a G-graded G-acted algebra

We consider the notations given in Section $\S 2$. We recall the definition of a G-graded G-acted algebra and an example of a G-graded G-acted algebra as in [9]:

Definition 1 An algebra $\mathscr{C}$ is a G-graded G-acted algebra if

1. $\mathscr{C}$ is G -graded, i.e. $\mathscr{C}=\oplus_{\mathrm{g} \in \mathrm{G}} \mathscr{C}_{\mathrm{g}}$;
2. G acts on $\mathscr{C}$ (always on the left in this article);
3. $\forall \mathrm{h} \in \mathrm{G}, \forall \mathrm{c} \in \mathscr{C}_{\mathrm{h}}$ we have that ${ }^{\mathrm{g}} \mathrm{c} \in \mathscr{C}_{\mathrm{gh}}$ for all $\mathrm{g} \in \mathrm{G}$.

We denote the identity component (the 1-component) of $\mathscr{C}$ by $\mathscr{Z}:=\mathscr{C}_{1}$, which is a G-acted algebra.

Let $A=\bigoplus_{g \in G} A_{g}$ be a strongly G-graded $\mathcal{O}$-algebra with identity component $B:=A_{1}$. For the sake of simplicity, we assume that $A$ is a crossed product (the generalization is not difficult, see for instance [7, §1.4.B.]). This means that we can choose invertible homogeneous elements $u_{g}$ in the component $A_{g}$.

Example 1 By Miyashita's theorem [7, p.22], we know that the centralizer $\mathrm{C}_{\mathrm{A}}(\mathrm{B})$ is a G-graded G -acted $\mathcal{O}$-algebra: for all $\mathrm{h} \in \mathrm{G}$ we have that

$$
C_{A}(B)_{h}=\left\{a \in A_{h} \mid a b=b a, \forall b \in B\right\}
$$

and the action is given by ${ }^{\mathrm{g}} \mathrm{c}=\mathrm{u}_{\mathrm{g}} \mathrm{cu}_{\mathrm{g}}^{-1}, \forall \mathrm{~g} \in \mathrm{G}, \forall \mathrm{c} \in \mathrm{C}_{\mathrm{A}}(\mathrm{B})$. Note that this definition does not depend on the choice of the elements $\mathfrak{u}_{\mathrm{g}}$ and that $\mathrm{C}_{\mathrm{A}}(\mathrm{B})_{1}=\mathrm{Z}(\mathrm{B})$ (the center of B$)$.

We recall the definition of a G-graded $\mathcal{O}$-algebra over a G-graded G-acted algebra $\mathscr{C}$ as in [9]:

Definition 2 Let $\mathscr{C}$ be a G-graded G-acted $\mathcal{O}$-algebra. We say that $\mathcal{A}$ is a Ggraded $\mathcal{O}$-algebra over $\mathscr{C}$ if there is a G-graded G-acted algebra homomorphism

$$
\zeta: \mathscr{C} \rightarrow \mathrm{C}_{\mathrm{A}}(\mathrm{~B})
$$

i.e. for any $\mathrm{h} \in \mathrm{G}$ and $\mathrm{c} \in \mathscr{C}_{\mathrm{h}}$, we have $\zeta(\mathrm{c}) \in \mathrm{C}_{\mathrm{A}}(\mathrm{B})_{\mathrm{h}}$, and for every $\mathrm{g} \in \mathrm{G}$, we have $\zeta\left({ }^{9} \mathrm{c}\right)={ }^{\mathrm{g}} \zeta(\mathrm{c})$.

An important example of a G-graded $\mathcal{O}$-algebra over a G-graded G-acted algebra, is given by the following lemma:

Lemma 1 Let P be a G-graded A-module. Assume that P is G-invariant. Let $\mathrm{A}^{\prime}=\operatorname{End}_{\mathcal{A}}(\mathrm{P})^{o p}$ be the set of all A -linear endomorphisms of P . Then $\mathrm{A}^{\prime}$ is a G-graded $\mathcal{O}$-algebra over $\mathrm{C}_{\mathrm{A}}(\mathrm{B})$.

Proof. By [3, Theorem 2.8], we have that there exists some $\mathrm{U} \in \mathrm{B}-\bmod$ such that P and $\mathrm{A} \otimes_{\mathrm{B}} \mathrm{U}$ are isomorphic as G -graded left A -modules, henceforth for simplicity we will identify $P$ as $A \otimes_{B} U$.

Because G is finite, E. C. Dade proved in [3, Corollary 3.10 and $\S 4]$ that $A^{\prime}=\operatorname{End}_{\mathcal{A}}(P)^{\text {op }}$ is a $G$-graded $\mathcal{O}$-algebra. Moreover, by $[3, \S 4]$ we have that P is actually a $G$-graded $\left(A, A^{\prime}\right)$-bimodule.

Now, the assumption that P is G-invariant, according to [3, Corollary 5.14] and $[7, \S 2.2 .1]$, implies that $A^{\prime}=\operatorname{End}_{\mathcal{A}}(P)^{\text {op }}$ is a crossed product and that $P$ is isomorphic to its g -suspension, $\mathrm{P}(\mathrm{g})$, for all $\mathrm{g} \in \mathrm{G}$. Henceforth, we can choose invertible homogeneous elements $\mathfrak{u}_{g}^{\prime}$ in the component $A_{g}^{\prime}$, for all $g \in G$ such that

$$
\mathrm{u}_{\mathrm{g}}^{\prime}: \mathrm{P} \rightarrow \mathrm{P}(\mathrm{~g}) .
$$

By taking the truncation functor $(-)_{1}$ (more details are given in [3]) we obtain the isomorphism:

$$
\left(u_{g}^{\prime}\right)_{1}: P_{1} \rightarrow(P(g))_{1}
$$

where $P_{1}=B \otimes_{B} U \simeq U$ and $(P(g))_{1}=A_{g} \otimes_{B} U=u_{g} B \otimes_{B} U$. We fix arbitrary $a \in A$ and $u \in U$. We have:

$$
\mathfrak{u}_{g}^{\prime}\left(a \otimes_{B} \mathfrak{u}\right)=a u_{g}^{\prime}\left(1_{A} \otimes_{B} \mathfrak{u}\right)
$$

but $1_{A} \otimes_{B} u \in P_{1}$, henceforth:

$$
u_{g}^{\prime}\left(a \otimes_{B} u\right)=a\left(u_{g}^{\prime}\right)_{1}\left(1_{A} \otimes_{B} u\right)
$$

but there exists an unique $b \in B$ such that $\left(u_{g}^{\prime}\right)_{1}\left(1_{A} \otimes_{B} u\right)=u_{g} b \otimes_{B} u=$ $u_{g} \otimes_{\mathrm{B}} \mathrm{bu}$. Therefore, by defining $\varphi_{\mathrm{g}}(\mathfrak{u}):=\mathrm{bu}$, we obtain a map $\varphi_{g}: \mathrm{U} \rightarrow \mathrm{U}$, which is clearly well-defined. Moreover, we have:

$$
u_{\mathrm{g}}^{\prime}\left(\mathrm{a} \otimes_{\mathrm{B}} \mathfrak{u}\right)=\mathrm{au}_{\mathrm{g}} \otimes_{\mathrm{B}} \varphi_{\mathrm{g}}(\mathfrak{u}), \text { for all } \mathrm{a} \in A \text { and } \mathfrak{u} \in \mathrm{U} .
$$

It is straightforward to prove that $\varphi_{\mathrm{g}}: \mathrm{U} \rightarrow \mathrm{U}$ admits an inverse and that

$$
\mathfrak{u}_{\mathrm{g}}^{\prime-1}\left(\mathrm{a} \otimes_{\mathrm{B}} \mathfrak{u}\right)=\mathfrak{a u}_{\mathrm{g}}^{-1} \otimes_{\mathrm{B}} \varphi_{\mathrm{g}}^{-1}(\mathfrak{u}), \text { for all } \mathrm{a} \in \mathcal{A} \text { and } \mathfrak{u} \in \mathrm{u} .
$$

We consider the G-graded algebra homomorphism from [8, Lemma 3.2.]:

$$
\theta: C_{A}(B) \rightarrow A^{\prime}=\operatorname{End}_{A}(P)^{o p}, \quad \theta(c)(a \otimes u)=a c \otimes u
$$

where $c \in C_{A}(B), a \in A$, and $u \in U$. First, we will prove that the image of $\theta$ is a subset of $C_{A^{\prime}}\left(B^{\prime}\right)$. Indeed, consider $b^{\prime} \in B^{\prime}$ and $c \in C_{A}(B)$. We want:

$$
\theta(c) \circ b^{\prime}=b^{\prime} \circ \theta(c) .
$$

Consider $\mathfrak{a} \otimes u \in A \otimes U=P$. We have:

$$
\left(\theta(c) \circ b^{\prime}\right)(a \otimes u)=\theta(c)\left(b^{\prime}(a \otimes u)\right)=a \theta(c)\left(b^{\prime}\left(1_{A} \otimes u\right)\right)
$$

because $b^{\prime}$ and $\theta(c)$ are $A$-linear. We fix $b^{\prime}\left(1_{A} \otimes u\right)=a_{0} \otimes u_{0} \in A \otimes_{B} U$, but because $\mathrm{b}^{\prime} \in \mathrm{B}^{\prime}=A_{1}^{\prime}=\operatorname{End}_{A}(\mathrm{P})_{1}^{\mathrm{op}}$ we know that $\mathrm{b}^{\prime}$ preserves the grading, so $1_{A} \in A_{1}$ implies that $a_{o} \in B$. Hence:

$$
\left(\theta(c) \circ b^{\prime}\right)(a \otimes u)=a \theta(c)\left(a_{0} \otimes u_{0}\right)=a a_{0} c \otimes u_{0}=a c a_{0} \otimes u_{0}
$$

Following, we have that:

$$
\begin{aligned}
\left(b^{\prime} \circ \theta(c)\right)(a \otimes u) & =\left(b^{\prime}(\theta(c)(a \otimes u))=b^{\prime}(a c \otimes u)\right. \\
& =\operatorname{acb}^{\prime}\left(1_{A} \otimes u\right)=a c a_{0} \otimes u_{0}
\end{aligned}
$$

Henceforth, the image of $\theta$ is a subset of $C_{A^{\prime}}\left(B^{\prime}\right)$. Second, we prove that $\theta$ is G-acted, in the sense that:

$$
\theta\left({ }^{9} c\right)={ }^{g}(\theta(c)), \text { for all } g \in G .
$$

Indeed, we fix $g \in G$, and $a \otimes u \in A \otimes_{B} U=P$. We have:

$$
\theta\left({ }^{g} c\right)(a \otimes u)=\theta\left(u_{g} c u_{g}^{-1}\right)(a \otimes u)=a u_{g} c u_{g}^{-1} \otimes u
$$

and

$$
\begin{aligned}
{ }^{g}(\theta(c))(a \otimes u) & =\left(u_{g}^{\prime} \cdot \theta(c) \cdot u_{g}^{\prime-1}\right)(a \otimes u)=\left(u_{g}^{\prime-1} \circ \theta(c) \circ u_{g}^{\prime}\right)(a \otimes u) \\
& =u_{g}^{\prime-1}\left(\theta(c)\left(u_{g}^{\prime}(a \otimes u)\right)\right)=u_{g}^{\prime-1}\left(\theta(c)\left(a u_{g} \otimes \varphi_{g}(u)\right)\right) \\
& =u_{g}^{\prime-1}\left(a u_{g} c \otimes \varphi_{g}(u)\right)=a u_{g} c u_{g}^{-1} \otimes \varphi_{g}^{-1}\left(\varphi_{g}(u)\right) \\
& =a u_{g} c u_{g}^{-1} \otimes u .
\end{aligned}
$$

Finally, by taking $\zeta^{\prime}: C_{A}(B) \rightarrow C_{A^{\prime}}\left(B^{\prime}\right)$ to be the corestriction of $\theta$ to $C_{A^{\prime}}\left(B^{\prime}\right)$, we obtain that $A^{\prime}$ is G-graded $\mathcal{O}$-algebra over $C_{A}(B)$, via the Ggraded G-acted homomorphism $\zeta^{\prime}$.

Let $A^{\prime}=\bigoplus_{g \in G} A_{g}^{\prime}$ be another strongly $G$-graded $\mathcal{O}$-algebra with the identity component $B^{\prime}:=A_{1}^{\prime}$. Again, we will consider that also $A^{\prime}$ is a crossed product, hence we will choose invertible homogeneous elements $\mathfrak{u}_{\mathfrak{g}}^{\prime}$ in the component $A_{g}^{\prime}$, for all $g \in G$.

Now, we assume that $A$ and $A^{\prime}$ are both strongly G-graded $\mathcal{O}$-algebras over a G-graded G-acted algebra $\mathscr{C}$, each endowed with G-graded G-acted algebra homomorphism $\zeta: \mathscr{C} \rightarrow \mathrm{C}_{A}(\mathrm{~B})$ and $\zeta^{\prime}: \mathscr{C} \rightarrow \mathrm{C}_{\mathcal{A}^{\prime}}\left(\mathrm{B}^{\prime}\right)$ respectively.

We recall the definition of a G-graded bimodule over $\mathscr{C}$ as in [9]:
Definition 3 We say that $\tilde{M}$ is a G-graded $\left(\mathcal{A}, \mathcal{A}^{\prime}\right)$-bimodule over $\mathscr{C}$ if:

1. $\tilde{M}$ is an $\left(\mathcal{A}, \mathcal{A}^{\prime}\right)$-bimodule;
2. $\tilde{\mathcal{M}}$ has a decomposition $\tilde{\mathcal{M}}=\bigoplus_{g \in G} \tilde{M}_{g}$ such that $A_{g} \tilde{\mathcal{M}}_{x} A_{h}^{\prime} \subseteq \tilde{\mathcal{M}}_{g \times h}$, for all $\mathrm{g}, \mathrm{x}, \mathrm{h} \in \mathrm{G}$;
3. $\tilde{\mathfrak{m}}_{\mathrm{g}} \cdot \mathrm{c}={ }^{\mathrm{g}} \mathrm{c} \cdot \tilde{\mathfrak{m}}_{\mathrm{g}}$, for all $\mathrm{c} \in \mathscr{C}, \tilde{\mathfrak{m}}_{\mathrm{g}} \in \tilde{\mathrm{M}}_{\mathrm{g}}, \mathrm{g} \in \mathrm{G}$, where $\mathrm{c} \cdot \tilde{\mathfrak{m}}=\zeta(\mathrm{c}) \cdot \tilde{\mathfrak{m}}$ and $\tilde{m} \cdot \mathrm{c}=\tilde{m} \cdot \zeta^{\prime}(\mathrm{c})$, for all $\mathrm{c} \in \mathscr{C}, \tilde{m} \in M$.

Remark 1 Condition 3. of Definition 3 can be rewritten (see [9] for the proof) as follows:

$$
\text { 3. } \mathrm{m} \cdot \mathrm{c}=\mathrm{c} \cdot \mathrm{~m} \text {, for all } \mathrm{c} \in \mathscr{C}, \mathrm{~m} \in \tilde{\mathrm{M}}_{1} .
$$

An example of a G-graded bimodule over a G-graded G-acted algebra is given by the following proposition:

Proposition 1 Let $\mathscr{C}$ be a G-graded G-acted algebra and $\mathcal{A}$ a strongly Ggraded $\mathcal{O}$-algebra over $\mathscr{C}$. Let P be a G-invariant G-graded A-module. Let $\mathrm{A}^{\prime}=\operatorname{End}_{\boldsymbol{A}}(\mathrm{P})^{o p}$. Then the following statements hold:

1. $\boldsymbol{A}^{\prime}$ is a G -graded $\mathcal{O}$-algebra over $\mathscr{C}$;
2. P is a G -graded $\left(\mathrm{A}, \mathrm{A}^{\prime}\right)$-bimodule over $\mathscr{C}$.

Proof. 1. By Lemma 1, we know that $A^{\prime}$ is $G$-graded $\mathcal{O}$-algebra over $C_{A}(B)$ and let $\theta: C_{A}(B) \rightarrow C_{A^{\prime}}\left(B^{\prime}\right)$ to be its G-graded G-acted structure homomorphism. Now, given that $\mathcal{A}$ is a strongly G -graded $\mathcal{O}$-algebra over $\mathscr{C}$, we have a G-graded G-acted algebra homomorphism $\zeta: \mathscr{C} \rightarrow \mathrm{C}_{A}(\mathrm{~B})$ and by taking $\zeta^{\prime}: \mathscr{C} \rightarrow \mathrm{C}_{\mathrm{A}^{\prime}}\left(\mathrm{B}^{\prime}\right)$ to be the G -graded G -acted algebra homomorphism obtained by composing $\zeta$ with $\theta$, we obtain that $A^{\prime}$ is also a G-graded $\mathcal{O}$ algebra over $\mathscr{C}$, with its structure given by $\zeta^{\prime}$. Hence, the first statement of this proposition was proved.
2. Without any loss in generality, we will identify $P$ with $A \otimes_{B} U$, for some $\mathrm{U} \in \mathrm{B}-$ mod. Following the proof of Lemma 1 we know that P is a G-graded ( $A, A^{\prime}$ )-bimodule. We now check that $P$ is G-graded $\left(A, A^{\prime}\right)$-bimodule over $\mathscr{C}$. Indeed, we fix $g \in G, p_{g}=a_{g} \otimes u \in P_{g}$ and $c \in \mathscr{C}$. We have:

$$
p_{g} \cdot c=\left(a_{g} \otimes u\right) \cdot c=\left(a_{g} \otimes u\right) \cdot \zeta^{\prime}(c)=a_{g} \zeta(c) \otimes u
$$

but $a_{g} \in A_{g}$ so there exists $a b \in B$ such that $a_{g}=u_{g} b$, therefore:

$$
\begin{aligned}
p_{g} \cdot c & =u_{g} b \zeta(c) \otimes u=u_{g} \zeta(c) b \otimes u \\
& =u_{g} \zeta(c) u_{g}^{-1} u_{g} b \otimes u={ }^{g} \zeta(c) a_{g} \otimes u \\
& =\zeta\left({ }^{g} c\right) a_{g} \otimes u=\zeta\left({ }^{g} c\right)\left(a_{g} \otimes u\right) \\
& =\zeta\left({ }^{g} c\right) \cdot p_{g}={ }^{g^{c}} \cdot{ }^{c} \cdot p_{g} .
\end{aligned}
$$

Henceforth, the last statement of this proposition has been proved.

## 4 Graded Morita contexts over $\mathscr{C}$

We consider the notations given in Section $\S 2$. Let $\mathscr{C}$ be a G-graded G-acted algebra. We introduce the notion of a G-graded Morita context over $\mathscr{C}$, following the treatment given in $[5, \S 12]$.

Definition 4 Consider the following Morita context:

$$
\left(A, A^{\prime}, \tilde{M}, \tilde{M}^{\prime}, f, g\right) .
$$

We call it a G-graded Morita context over $\mathscr{C}$ if:

1. A and $\mathrm{A}^{\prime}$ are strongly G -graded $\mathcal{O}$-algebras over $\mathscr{C}$;
2. ${ }_{A} \tilde{M}_{A^{\prime}}$ and ${ }_{A^{\prime}}, \tilde{M}_{A}^{\prime}$ are G-graded bimodules over $\mathscr{C}$;
3. $\mathrm{f}: \tilde{\mathrm{M}} \otimes_{\mathcal{A}^{\prime}} \tilde{\mathrm{M}}^{\prime} \rightarrow \mathrm{A}$ and $\mathrm{g}: \tilde{\mathrm{M}}^{\prime} \otimes_{\mathrm{A}} \tilde{M} \rightarrow \mathrm{~A}^{\prime}$ are G-graded bimodule homomorphisms such that by setting $\mathfrak{f}\left(\tilde{m} \otimes \tilde{m}^{\prime}\right)=\tilde{\mathfrak{m}} \tilde{m}^{\prime}$ and $\mathrm{g}\left(\tilde{m}^{\prime} \otimes \tilde{\mathfrak{m}}\right)=$ $\tilde{m}^{\prime} \tilde{m}$, we have the associative laws:

$$
\left(\tilde{m} \tilde{m}^{\prime}\right) \tilde{\mathfrak{m}}=\tilde{m}\left(\tilde{m}^{\prime} \tilde{\mathfrak{n}}\right) \quad \text { and } \quad\left(\tilde{m}^{\prime} \tilde{m}\right) \tilde{\mathfrak{n}}^{\prime}=\tilde{m}^{\prime}\left(\tilde{m} \tilde{\mathfrak{n}}^{\prime}\right)
$$

for all $\tilde{\mathrm{m}}, \tilde{\mathrm{n}} \in \tilde{M}, \tilde{\mathrm{~m}}^{\prime}, \tilde{\mathrm{n}}^{\prime} \in \tilde{M}^{\prime}$.
If f and g are isomorphisms, then $\left(\mathrm{A}, \mathrm{A}^{\prime}, \tilde{M}, \tilde{M}^{\prime}, \mathrm{f}, \mathrm{g}\right)$ is called a surjective G-graded Morita context over $\mathscr{C}$.

As an example of a G-graded Morita context over $\mathscr{C}$, we have the following proposition which arises from [5, Proposition 12.6].

Proposition 2 Let A be a strongly G-graded $\mathcal{O}$-algebra over $\mathscr{C}$, let P be a Ginvariant G -graded A -module, let $\mathrm{A}^{\prime}=\operatorname{End}_{\mathcal{A}}(\mathrm{P})^{o p}$ and let $\mathrm{P}^{*}:=\operatorname{Hom}_{\mathcal{A}}(\mathrm{P}, \mathrm{A})$ be the A -dual of P . Then

$$
\left(A, A^{\prime}, P, P^{*},(\cdot, \cdot),[\cdot, \cdot]\right)
$$

is a G-graded Morita context over $\mathscr{C}$, where $(\cdot, \cdot)$ is a G -graded $(\mathrm{A}, \mathrm{A})$-homomorphism, called the evaluation map, defined by:

$$
\begin{aligned}
& (\cdot, \cdot): \mathrm{P} \otimes_{\mathcal{A}^{\prime}} \mathrm{P}^{*} \rightarrow \mathrm{~A}, \\
& \mathrm{x} \otimes \varphi \mapsto \varphi(\mathrm{x}), \text { for all } \varphi \in \mathrm{P}^{*}, \mathrm{x} \in \mathrm{P},
\end{aligned}
$$

and where $[\cdot, \cdot]$ is a G -graded $\left(\mathrm{A}^{\prime}, \mathrm{A}^{\prime}\right)$-homomorphism defined by:

$$
\begin{aligned}
& {[\cdot, \cdot]: \mathrm{P}^{*} \otimes_{\mathrm{A}} \mathrm{P} \rightarrow \mathrm{~A}^{\prime},} \\
& \varphi \otimes \mathrm{x} \mapsto[\varphi, x], \text { for all } \varphi \in \mathrm{P}^{*}, x \in \mathrm{P},
\end{aligned}
$$

where for every $\varphi \in \mathrm{P}^{*}$ and $\mathrm{x} \in \mathrm{P},[\varphi, \chi]$ is an element of $\mathrm{A}^{\prime}$ such that

$$
y[\varphi, x]=\varphi(y) \cdot x, \text { for all } y \in P
$$

Proof. For the sake of simplicity, we will assume that $\mathcal{A}$ is a crossed product as in Section $\S 3$. By Proposition 1, we have that $A^{\prime}$ is also a $G$-graded $\mathcal{O}$-algebra over $\mathscr{C}$ and that $P$ is a $G$-graded $\left(A, A^{\prime}\right)$-bimodule over $\mathscr{C}$. Now, it is known that the $A$-dual of $P, P^{*}:=\operatorname{Hom}_{\mathcal{A}}(P, A)$ is a $\left(A^{\prime}, A\right)$-bimodule, where for each $\varphi \in P^{*}$ and for each $p \in P$, we have:

$$
\left(\mathfrak{a}^{\prime} \varphi \mathrm{a}\right)(\mathrm{p})=\left(\varphi\left(\mathrm{pa}^{\prime}\right)\right) \mathrm{a}
$$

for all $a^{\prime} \in A^{\prime}$ and $a \in A$. By [7, §1.6.4.], we know that $P^{*}$ is actually a $G$-graded $\left(A^{\prime}, A\right)$-bimodule, where for all $g \in G$, the $g$-component is defined as follows:

$$
P_{g}^{*}=\left\{\varphi \in P^{*} \mid \varphi\left(P_{x}\right) \subseteq A_{x g}, \text { for all } x \in G\right\} .
$$

We prove that $P^{*}$ is a $G$-graded ( $A^{\prime}, A$ )-bimodule over $\mathscr{C}$. Consider $g, h \in G$, $\varphi_{g} \in P_{g}^{*}, c \in C$ and $p_{h} \in P_{h}$. We have:

$$
\left(\varphi_{g} \mathfrak{c}\right)\left(p_{h}\right)=\left(\varphi_{g}\right)\left(p_{h}\right) c .
$$

Because $\left(\varphi_{g}\right)\left(p_{h}\right) \in A_{h g}$ we can choose a homogeneous element $u_{h g} \in A_{\mathrm{hg}}$ and $b \in B$ such that $\left(\varphi_{g}\right)\left(p_{h}\right)=u_{h g} b$. Henceforth,

$$
\begin{aligned}
\left(\varphi_{g} c\right)\left(p_{h}\right) & =u_{h g} b c=u_{h g} c b=u_{h g} c u_{h g}^{-1} u_{h g} b={ }^{h g} c u_{h g} b \\
& ={ }^{h g} c\left(\varphi_{g}\right)\left(p_{h}\right)=\left(\varphi_{g}\right)\left({ }^{h g} c p_{h}\right) \\
& =\left(\varphi_{g}\right)\left(p_{h}{ }^{g} c\right)=\left({ }^{g} c \varphi_{g}\right)\left(p_{h}\right)
\end{aligned}
$$

thus $\varphi_{g} \mathrm{c}={ }^{9} \mathrm{C} \varphi_{g}$, therefore $\mathrm{P}^{*}$ is a G-graded $\left(\mathcal{A}^{\prime}, A\right)$-bimodule over $\mathscr{C}$. Next, following $[5, \S 12]$, it is clear that $(\cdot, \cdot)$ and $[\cdot, \cdot]$ are an $(A, A)$-homomorphism and an $\left(A^{\prime}, A^{\prime}\right)$-homomorphism, respectively. We now check if they are graded, as in the sense of $[2, \S 3]$ : Indeed, consider $p_{g} \in P_{g}$ and $\varphi_{h} \in P_{h}^{*}$. We have:

$$
\left(p_{g}, \varphi_{h}\right)=\varphi_{h}\left(p_{g}\right)
$$

which is an element of $A_{g h}$, given the gradation of $P^{*}:=\operatorname{Hom}_{\mathcal{A}}(P, A)$. Also, for every $y_{k} \in P_{k}$, we have:

$$
\left[\varphi_{h}, p_{g}\right]\left(y_{k}\right)=\varphi_{h}\left(y_{k}\right) p_{g}
$$

which is an element of $A_{\text {khg }}$, given the gradation of $\mathrm{P}^{*}$ and of P , therefore [ $\varphi_{h}, p_{g}$ ] is an element of $\mathcal{A}_{h g}$. Finally, we verify the associative law of the two homomorphism: Let $p, q \in P$ and $\varphi, \psi \in P^{*}$. We have:

$$
(p, \varphi) q=\varphi(p) q \quad \text { and } \quad p[\varphi, q]=\varphi(p) q
$$

hence

$$
(p, \varphi) q=p[\varphi, q]
$$

and for all $y \in P$, we have:

$$
([\varphi, p] \psi)(y)=\psi(y[\varphi, p])=\psi(\varphi(y) p)=\varphi(y) \psi(p)
$$

because $\varphi(\mathrm{y}) \in \mathcal{A}$ and $\psi$ is $A$-linear, and also we have

$$
(\varphi(p, \psi))(y)=(\varphi \psi(p))(y)=\varphi(y) \psi(p)
$$

because $\psi(p) \in A$, hence

$$
[\varphi, p] \psi=\varphi(p, \psi)
$$

Therefore $\left(A, A^{\prime}, P, P^{*},(\cdot, \cdot),[\cdot, \cdot]\right)$ is a G-graded Morita context over $\mathscr{C}$.
If $\left(A, A^{\prime}, \tilde{M}, \tilde{M}^{\prime}, f, g\right)$ is a surjective G-graded Morita context over $\mathscr{C}$, then by Proposition 12.7 of [5], we have that $A^{\prime}$ is isomorphic to $\operatorname{End}_{\mathcal{A}}(\tilde{M})^{\text {op }}$ and we have a bimodule isomorphism between $\tilde{M}^{\prime}$ and $\tilde{M}^{*}=\operatorname{Hom}_{\mathcal{A}}(\tilde{M}, A)$. Henceforth, in this situation, the example given by Proposition 2 is essentially unique up to an isomorphism.

Given Corollary 12.8 of [5], the example given by Proposition 2 is a surjective G-graded Morita context over $\mathscr{C}$ if and only if ${ }_{A} \mathrm{P}$ is a progenerator.

## 5 Graded Morita theorems over $\mathscr{C}$

Consider the notations given in $\S 2$. Let $\mathscr{C}$ be a G-graded G-acted algebra. We denote by $A$ and $A^{\prime}$ two strongly $G$-graded $\mathcal{O}$-algebras over $\mathscr{C}$ (with identity components $\mathrm{B}:=A_{1}$ and $\mathrm{B}^{\prime}:=A_{1}^{\prime}$ ), each endowed with G-graded G-acted algebra homomorphism $\zeta: \mathscr{C} \rightarrow \mathrm{C}_{A}(\mathrm{~B})$ and $\zeta^{\prime}: \mathscr{C} \rightarrow \mathrm{C}_{\mathcal{A}^{\prime}}\left(\mathrm{B}^{\prime}\right)$ respectively. According to [4] we have the following definitions:

Definition 5 1. We say that the functor $\tilde{\mathcal{F}}: \mathrm{A}-\mathrm{Gr} \rightarrow \mathrm{A}^{\prime}$ - Gr is G -graded if for every $\mathrm{g} \in \mathrm{G}, \tilde{\mathcal{F}}$ commutes with the g -suspension functor, i.e. $\tilde{\mathcal{F}} \circ \mathrm{T}_{\mathrm{g}}^{\mathrm{A}}$ is naturally isomorphic to $\mathrm{T}_{\mathrm{g}}^{\mathrm{A}^{\prime}} \circ \tilde{\mathcal{F}}$;
2. We say that $\mathcal{A}$ and $\mathcal{A}^{\prime}$ are G-graded Morita equivalent if there is a Ggraded equivalence: $\tilde{\mathcal{F}}: \mathrm{A}-\mathrm{Gr} \rightarrow \mathrm{A}^{\prime}$-Gr.

Assume that $A$ and $A^{\prime}$ are G-graded Morita equivalent. Therefore, we can consider the G -graded functors:

which give a G-graded Morita equivalence between $A$ and $A^{\prime}$. By Gordon and Green's result [4, Corollary 10], this is equivalent to the existence of a Morita equivalence between $A$ and $A^{\prime}$ given by the following functors:

$$
\text { A-Mod } \underset{\mathcal{G}}{\rightleftarrows} A^{\prime}-\operatorname{Mod} ;
$$

such that the following diagram is commutative:

in the sense that:

$$
\mathrm{U}^{\prime} \circ \tilde{\mathrm{F}}=\mathrm{F} \circ \mathrm{U}, \quad \mathrm{U} \circ \tilde{\mathrm{G}}=\mathrm{G} \circ \mathrm{U}^{\prime},
$$

where $\mathrm{U}^{\prime}$ is the forgetful functor from $\mathrm{A}^{\prime}-\mathrm{Gr}$ to $\mathrm{A}^{\prime}$-Mod.

Lemma 2 If $\tilde{\mathrm{P}}$ is a G-graded A-module, then $\tilde{\mathrm{P}}$ and $\tilde{\mathcal{F}}(\tilde{\mathrm{P}})$ have the same stabilizer in G.

Proof. Let $g \in G_{\tilde{p}}$. We have $\tilde{P} \simeq \tilde{P}(g)$ as $G$-graded $A$-modules. Because $\tilde{\mathcal{F}}$ is a graded functor, we have that it commutes with the $g$-suspension functor. Thus $\tilde{\mathcal{F}}(\tilde{\mathrm{P}}(\mathrm{g})) \simeq \tilde{\mathcal{F}}(\tilde{\mathrm{P}})(\mathrm{g})$ in $A^{\prime}$-Gr. Henceforth, $\tilde{\mathcal{F}}(\tilde{\mathrm{P}}) \simeq \tilde{\mathcal{F}}(\tilde{\mathrm{P}})(\mathrm{g})$ in $A^{\prime}$-Gr, thus $g \in G_{\tilde{\mathcal{F}}(\tilde{P})}$. Hence $G_{\tilde{p}} \subseteq G_{\tilde{\mathcal{F}}(\tilde{P})}$. The converse, $G_{\tilde{\mathcal{F}}(\tilde{P})} \subseteq G_{\tilde{p}}$, is straightforward, thus $G_{\tilde{P}}=G_{\tilde{\mathcal{F}}(\tilde{P})}$.

Consider $\tilde{\mathrm{P}}$ and $\tilde{\mathrm{Q}}$ two G-graded A -modules. We have the following morphism:

$$
\begin{equation*}
\operatorname{Hom}_{\mathcal{A}}(\tilde{\mathrm{P}}, \tilde{\mathrm{Q}}) \longrightarrow \operatorname{Hom}_{\mathcal{A}^{\prime}}(\tilde{\mathcal{F}}(\tilde{\mathrm{P}}), \tilde{\mathcal{F}}(\tilde{\mathrm{Q}})) \tag{*}
\end{equation*}
$$

By following the proofs of Lemma 1 and Proposition 1, we have a G-graded homomorphism from $\mathscr{C}$ to $\operatorname{End}_{A}(\tilde{P})^{\text {op }}$ (the composition between the structure homomorphism $\zeta: \mathscr{C} \rightarrow C_{A}(B)$ and the morphism $\theta: C_{A}(B) \rightarrow \operatorname{End}_{A}(\tilde{P})^{\text {op }}$ from $[8, \operatorname{Lemma} 3.2]$.$) and that \tilde{P}$ is a $G$-graded $\left(A, \operatorname{End}_{A}(\tilde{P})^{\mathrm{op}}\right)$-bimodule. Then, by the restriction of scalars we obtain that $\tilde{\mathrm{P}}$ is a right $\mathscr{C}$-module. Analogously $\tilde{\mathrm{Q}}, \tilde{\mathcal{F}}(\tilde{\mathrm{P}})$ and $\tilde{\mathcal{F}}(\tilde{\mathrm{Q}})$ are also right $\mathscr{C}$-modules, thus $\operatorname{Hom}_{\mathcal{A}}(\tilde{\mathrm{P}}, \tilde{\mathrm{Q}})$ and $\operatorname{Hom}_{\mathcal{A}}(\tilde{\mathcal{F}}(\tilde{\mathrm{P}}), \tilde{\mathcal{F}}(\tilde{\mathrm{Q}}))$ are G-graded $(\mathscr{C}, \mathscr{C})$-bimodules. This allows us to state the following definition:

Definition 6 1. We say that the functor $\tilde{\mathcal{F}}$ is over $\mathscr{C}$ if the morphism $\tilde{\mathcal{F}}$ $($ see $(*))$ is a morphism of G-graded $(\mathscr{C}, \mathscr{C})$-bimodules;
2. We say that $A$ and $A^{\prime}$ are G-graded Morita equivalent over $\mathscr{C}$ if there is a G-graded equivalence over $\mathscr{C}: \tilde{\mathcal{F}}: A-G r \rightarrow A^{\prime}$-Gr.

Theorem 1 (Graded Morita I over $\mathscr{C})$ Let $\left(A, A^{\prime}, \tilde{M}, \tilde{M}^{\prime}, f, g\right)$ be a surjective G-graded Morita context over $\mathscr{C}$. Then the functors

$$
\begin{aligned}
& \tilde{M}^{\prime} \otimes_{A}-: A-\mathrm{Gr} \rightarrow A^{\prime}-\mathrm{Gr} \\
& \tilde{M} \otimes_{A^{\prime}}-: A^{\prime}-\mathrm{Gr} \rightarrow A-\mathrm{Gr}
\end{aligned}
$$

are inverse G-graded equivalences over $\mathscr{C}$.
Proof. Given [2, Theorem 3.2 (Graded Morita I) 6.] we already know that the pair of functors $\tilde{M}^{\prime} \otimes_{A}-$ and $\tilde{M} \otimes_{A^{\prime}}-$ are inverse G-graded equivalences. It remains to prove that they are also over $\mathscr{C}$. We will only prove that the functor $\tilde{M}_{\tilde{P}}^{\prime} \otimes_{A}-$ is over $\mathscr{C}$ because the proof for the latter functor is similar. Consider $\tilde{\mathrm{P}}$ and $\tilde{\mathrm{Q}}$ two G-graded $A$-modules.

First, we will prove that the morphism

$$
\begin{equation*}
\tilde{M}^{\prime} \otimes_{A}-: \operatorname{Hom}_{\mathcal{A}}(\tilde{\mathrm{P}}, \tilde{\mathrm{Q}}) \rightarrow \operatorname{Hom}_{\mathcal{A}^{\prime}}\left(\tilde{M}^{\prime} \otimes_{\mathrm{A}} \tilde{\mathrm{P}}, \tilde{M}^{\prime} \otimes_{A} \tilde{\mathrm{Q}}\right) \tag{**}
\end{equation*}
$$

is a $(\mathscr{C}, \mathscr{C})$-bimodule homomorphism. Indeed, consider $\varphi \in \operatorname{Hom}_{\mathcal{A}}(\tilde{P}, \tilde{Q})$, then $\tilde{M}^{\prime} \otimes_{A} \varphi \in \operatorname{Hom}_{\mathcal{A}^{\prime}}\left(\tilde{M}^{\prime} \otimes_{A} \tilde{P}, \tilde{M}^{\prime} \otimes_{A} \tilde{Q}\right)$. Consider $c, c^{\prime} \in \mathscr{C}$. We only need to prove that

$$
\tilde{M}^{\prime} \otimes_{\mathrm{A}}\left(c \varphi c^{\prime}\right)=c\left(\tilde{M}^{\prime} \otimes_{\mathrm{A}} \varphi\right) c^{\prime}
$$

Let $\tilde{m}^{\prime} \in \tilde{M}^{\prime}$ and $\tilde{\mathrm{p}} \in \tilde{\mathrm{P}}$. We have:

$$
\left(\tilde{M}^{\prime} \otimes_{\mathrm{A}}\left(c \varphi c^{\prime}\right)\right)\left(\tilde{m}^{\prime} \otimes \tilde{\mathfrak{p}}\right)=\tilde{\mathfrak{m}}^{\prime} \otimes\left(c \varphi c^{\prime}\right)(\tilde{\mathfrak{p}})=\tilde{\mathfrak{m}}^{\prime} \otimes \varphi(\tilde{p} c) \mathfrak{c}^{\prime}
$$

and

$$
\begin{aligned}
\left(c\left(\tilde{M}^{\prime} \otimes_{A} \varphi\right) c^{\prime}\right)\left(\tilde{m}^{\prime} \otimes \tilde{p}\right) & =\left(\left(\tilde{M}^{\prime} \otimes_{A} \varphi\right)\left(\left(\tilde{m}^{\prime} \otimes \tilde{p}\right) c\right)\right) c^{\prime} \\
& =\left(\left(\tilde{M}^{\prime} \otimes_{A} \varphi\right)\left(\tilde{m}^{\prime} \otimes \tilde{p} c\right)\right) c^{\prime} \\
& =\left(\tilde{m}^{\prime} \otimes \varphi(\tilde{p} c)\right) c^{\prime} \\
& =\tilde{m}^{\prime} \otimes \varphi(\tilde{p} c) c^{\prime}
\end{aligned}
$$

Henceforth $\tilde{M}^{\prime} \otimes_{A}\left(c \varphi c^{\prime}\right)=c\left(\tilde{M}^{\prime} \otimes_{A} \varphi\right) c^{\prime}$, thus the morphism $\tilde{M}^{\prime} \otimes_{A}-($ see $(* *))$ is a $(\mathscr{C}, \mathscr{C})$-bimodule homomorphism.

Second, we will prove that the morphism $\tilde{M}^{\prime} \otimes_{\mathrm{A}}-$ is a G-graded $(\mathscr{C}, \mathscr{C})$ bimodule homomorphism, i.e. it is grade preserving. Consider $g \in G$ and $\varphi_{g} \in \operatorname{Hom}_{\mathcal{A}}(\tilde{P}, \tilde{Q})_{g}$. We want $\tilde{M}^{\prime} \otimes_{A} \varphi_{g} \in \operatorname{Hom}_{\mathcal{A}^{\prime}}\left(\tilde{M}_{\tilde{P}}^{\prime} \otimes_{A} \tilde{P}, \tilde{M}^{\prime} \otimes_{A} \tilde{Q}\right)_{g}$, i.e. by $[7, \S 1.2]$, if for some $h \in G$ and $\tilde{m}^{\prime} \otimes \tilde{p} \in\left(\tilde{M}^{\prime} \otimes_{A} \tilde{P}\right)_{h}$, then we must have $\left(\tilde{M}^{\prime} \otimes_{A} \varphi_{g}\right)\left(\tilde{m}^{\prime} \otimes \tilde{p}\right) \in\left(\tilde{M}^{\prime} \otimes_{A} \tilde{Q}\right)_{h g}$. Beforehand, because $\tilde{m}^{\prime} \otimes \tilde{p} \in\left(\tilde{M}^{\prime} \otimes_{A} \tilde{P}\right)_{h}$, by $[7, \S 1.6 .4]$, there exists some $x, y \in G$ with $h=x y$ such that $\tilde{m}^{\prime} \in \tilde{M}_{x}^{\prime}$ and $\tilde{p} \in \tilde{P}_{y}$. We have:

$$
\begin{aligned}
\left(\tilde{M}^{\prime} \otimes_{A} \varphi_{g}\right)\left(\tilde{m}^{\prime} \otimes \tilde{p}\right) & =\tilde{m}^{\prime} \otimes_{g}(\tilde{p}) \in \tilde{M}_{\chi}^{\prime} \otimes_{A} \varphi_{g}\left(\tilde{P}_{y}\right) \\
& \subseteq \tilde{M}_{x}^{\prime} \otimes_{A} \tilde{Q}_{y g} \subseteq\left(\tilde{M}^{\prime} \otimes_{A} \tilde{Q}\right)_{x y g} \\
& =\left(\tilde{M}^{\prime} \otimes_{A} \tilde{Q}\right)_{h g} .
\end{aligned}
$$

Henceforth, the morphism $\tilde{M}^{\prime} \otimes_{A}-: \operatorname{Hom}_{\mathcal{A}}(\tilde{P}, \tilde{Q}) \rightarrow \operatorname{Hom}_{\mathcal{A}^{\prime}}\left(\tilde{M}^{\prime} \otimes_{\mathrm{A}} \tilde{P}, \tilde{M}^{\prime} \otimes_{\mathrm{A}}\right.$ $\tilde{\mathrm{Q}})$ is a G-graded ( $\mathscr{C}, \mathscr{C})$-bimodule homomorphism.

By Proposition 2 and the observations made in Section $\S 4$, the following corollary is straightforward.

Corollary 1 Let P be a G-invariant G-graded A -module and $\mathrm{A}^{\prime}=\operatorname{End}_{\mathcal{A}}(\mathrm{P})^{o p}$. If ${ }_{\mathrm{A}} \mathrm{P}$ is a progenerator, then $\mathrm{P} \otimes_{A^{\prime}}-$ is a G -graded Morita equivalence over $\mathscr{C}$ between $A^{\prime}-\mathrm{Gr}$ and $\mathrm{A}-\mathrm{Gr}$, with $\mathrm{P}^{*} \otimes_{\mathrm{A}}-$ as its inverse.

Theorem 2 (Graded Morita II over $\mathscr{C}$ ) Assume that $\mathcal{A}$ and $\mathcal{A}^{\prime}$ are Ggraded Morita equivalent over $\mathscr{C}$ and let

be inverse G-graded equivalences over $\mathscr{C}$. Then this equivalence is given by the following G -graded bimodules over $\mathscr{C}: \mathrm{P}=\tilde{\mathcal{F}}(\mathrm{A})$ and $\mathrm{Q}=\tilde{\mathcal{G}}\left(\mathrm{A}^{\prime}\right)$. More exactly, P is a G -graded $\left(\mathrm{A}^{\prime}, \mathrm{A}\right)$-bimodule over $\mathscr{C}, \mathrm{Q}$ is a G -graded $\left(\mathrm{A}, \mathrm{A}^{\prime}\right)$-bimodule over $\mathscr{C}$ and the following natural equivalences of functors hold:

$$
\tilde{\mathcal{F}} \simeq \mathrm{P} \otimes_{\mathcal{A}}-\quad \text { and } \quad \tilde{\mathcal{G}} \simeq \mathrm{Q} \otimes_{\mathcal{A}^{\prime}}-
$$

Proof. By [4, Corollary 10 (Gordon-Green)], we know that $\mathrm{P}=\tilde{\mathcal{F}}(\mathcal{A})$ is a $G$-graded ( $A^{\prime}, A$ )-bimodule, $Q=\tilde{\mathcal{G}}\left(A^{\prime}\right)$ is a $G$-graded ( $A, A^{\prime}$ )-bimodule and that that the following natural equivalences of functors hold: $\mathcal{F} \simeq \mathrm{P} \otimes_{A}-$ and $\tilde{\mathcal{G}} \simeq \mathrm{Q} \otimes_{\mathcal{A}^{\prime}}-$. Moreover, we have that ${ }_{A} \mathrm{P}$ is a progenerator.

It remains to prove that P and Q are G -graded bimodules over $\mathscr{C}$. We will only prove that P is G -graded bimodule over $\mathscr{C}$, because for Q the reasoning is similar.

By the hypothesis, $A$ and $A^{\prime}$ are G-graded Morita equivalent over $\mathscr{C}$, hence $\tilde{\mathcal{F}}$ and $\tilde{\mathcal{G}}$ are over $\mathscr{C}$. Therefore the function:

$$
\operatorname{Hom}_{\mathcal{A}}\left({ }_{A} A,{ }_{A} A\right) \xrightarrow{\tilde{\mathcal{F}}} \operatorname{Hom}_{\mathcal{A}^{\prime}}\left({ }_{A^{\prime}}, P,{ }_{A^{\prime}}, P\right)
$$

is an isomorphism of $G$-graded $(\mathscr{C}, \mathscr{C})$-bimodules, where if $f \in \operatorname{Hom}_{\mathcal{A}}\left({ }_{A} A,{ }_{A} A\right)$, we have that $\left(c_{1} f c_{2}\right)(a)=f\left(a c_{1}\right) c_{2}$, for all $a \in A, c_{1}, c_{2} \in \mathscr{C}$. This means that the function

$$
\alpha: A \rightarrow \operatorname{End}_{A^{\prime}}(P)^{\text {op }}, \quad \alpha(a)=\tilde{\mathcal{F}}(\rho(a)), \text { for all } a \in A,
$$

(where $\rho(c): a \mapsto a c$, for all $a \in A$ ) is an isomorphism of G-graded $(\mathscr{C}, \mathscr{C})$ bimodules. Moreover, by the bimodule structure definition of P (see [1]), we have that $\alpha(a)(p)=p a$ for all $a \in A$ and for all $p \in P$.

It is clear that ${ }_{A} \mathcal{A}$ is G-invariant, hence by Lemma 2, P is also G-invariant. Henceforth, by Proposition 1, P is a $G$-graded $\left(\mathrm{A}^{\prime}, \operatorname{End}_{A^{\prime}}(P)^{\text {op }}\right)$-bimodule over $\mathscr{C}$. Consider the structural homomorphisms $\zeta: \mathscr{C} \rightarrow A, \zeta^{\prime}: \mathscr{C} \rightarrow A^{\prime}$ and $\zeta^{\prime \prime}: \mathscr{C} \rightarrow \operatorname{End}_{\mathcal{A}^{\prime}}(P)^{\text {op }}$, thus for all $a \in A$ and for all $c_{1}, c_{2} \in \mathscr{C}$ we have:

$$
\alpha\left(\zeta\left(c_{1}\right) a \zeta\left(c_{2}\right)\right)=\zeta^{\prime \prime}\left(c_{1}\right) \alpha(a) \zeta^{\prime \prime}\left(c_{2}\right) .
$$

By taking $a=1_{\mathrm{A}}$ and $c_{2}=1_{\mathscr{C}}$ we obtain $\alpha \circ \zeta=\zeta^{\prime \prime}$.
Let $g \in G, p_{g} \in P_{g}$ and $c \in \mathscr{C}$. We want $p_{g} \cdot c={ }^{g} c \cdot p_{g}$. We have:

$$
p_{g} \cdot c=p_{g} \cdot \alpha(\zeta(c))=p_{g} \cdot \zeta^{\prime \prime}(c)={ }^{g} \zeta^{\prime}(c) p_{g}={ }^{9} c p_{g} .
$$

Henceforth the statement is proved.

## 6 Conclusion

We have developed a G-graded Morita theory over a G-graded G-acted algebra for the case of finite groups.

In Section §3, we recalled from [9] the notions of a G-graded G-acted algebra, of a G-graded algebra over a G-graded G-acted algebra and that of a G-graded bimodule over a G-graded G-acted algebra and we gave some useful examples for each notion.

In Section §4, we introduced the notion of a G-graded Morita context over a G-graded G-acted algebra and gave a standard example.

The main results are in Section $\S 5$, where a notion of graded functors over G-graded G-acted algebras and of graded Morita equivalences over G-graded G-acted algebras are introduced and two Morita-type theorems are proved using these notions: We proved that by taking a G-graded bimodule over a G-graded G-acted algebra we obtain a G-graded Morita equivalence over the said G-graded G-acted algebra and that by being given a G-graded Morita equivalence over a G-graded G-acted algebra, we obtain a G-graded bimodule over the said G-graded G-acted algebra, which induces the given G-graded Morita equivalence.

## References

[1] F. W. Anderson, K. R. Fuller, Rings and Categories of Modules, Graduate Texts in Mathematics, Vol. 13, 2nd Ed., Springer-Verlag, Berlin-Heidelberg-New York (1992), 55-265.
[2] P. Boisen, Graded Morita Theory, Journal of Algebra, 164 (1994), 1-25.
[3] E. C. Dade, Group-Graded Rings and Modules, Math. Z., 174 (1980), 241-262.
[4] A. del Río, Graded rings and equivalences of categories, Communications in Algebra, 19 (3) (1991), 997-1012.
[5] C. Faith, Algebra: Rings, Modules and Categories I, Springer-Verlag, Berlin-Heidelberg-New York (1973), 443-453.
[6] R. Gordon, E. L. Green, Graded Artin Algebras, J. of Algebra, 76 (1980), 241-262.
[7] A. Marcus, Representation theory of group-graded algebras, Nova Science (1999), 22-162.
[8] A. Marcus, V. A. Minuță, Group graded endomorphism algebras and Morita equivalences, Mathematica, 62 (85), $\mathrm{N}^{\mathrm{o}} 1$ (2020),
DOI:10.24193/mathcluj.2020.1.08, 73-80.
[9] A. Marcus, V. A. Minuță, Character triples and equivalences over a group graded G-algebra, arXiv:1912.05666, preprint (2019), 1-23.
[10] C. Menini, C. Năstăsescu, When is R-gr Equivalent to the Category of Modules?, J. of Pure and Appl. Algebra, 51 (1988), 277-291.
[11] B. Späth, Reduction theorems for some global-local conjectures, Local Representation Theory and Simple Groups, European Mathematical Society (2018), 23-61.

# Fixed points for a pair of weakly compatible mappings satisfying a new type of $\phi$ - implicit relation in $S$ - metric spaces 

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#### Abstract

The purpose of this paper is to introduce a new type of $\phi$ implicit relation in $S$ - metric spaces and to prove a general fixed point for a pair of weakly compatible mappings, which generalize Theorems 1, 2, 4 [23], Theorems 1-7 [13], Corollary 2.19 [13], Theorems 2.2, 2.4 [19], Theorems 3.2, 3.3, 3.4 [20] and other known results.


## 1 Introduction

Let $X$ be a nonempty set and $f, g: X \rightarrow X$ two self mappings. A point $x \in X$ is said to be a coincidence point of $f$ and $g$ if $f x=g x=w$. The set of all coincidence points of $f$ and $g$ is denoted $\mathcal{C}(f, g)$ and $w$ is said to be a point of coincidence of $f$ and $g$.

In [8], Jungck defined $f$ and $g$ to be weakly compatible if $f g x=g f x$, for all $x \in \mathcal{C}(f, g)$.

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The notion of weakly compatible mappings is used to proof the existence of common fixed point for pairs of mappings.

A new class of generalized metric space, named D - metric space, is introduced in $[5,6]$. In [11, 12], Mustafa and Sims proved that most of the claims concerning the fundamental topological structures on D - metric spaces are incorrect and introduced a new generalized metric spaces, named G - metric space. There exists a vast literature in the study of fixed points in G - metric spaces.

In [10], Mustafa initiated the study of fixed points for weakly compatible mappings in G-metric spaces.

Recently in [22], the authors introduced a new class of generalized metric space, named $S$ - metric space. Quite recently in [7], the authors proved that the notions of $G$ - metric spaces and $S$ - metric space are independent.

Other results in the study of fixed points in $S$ - metric space are obtained in $[13,19,20,21]$ and in other papers. Some results of fixed points for weakly compatible mappings in $S$ - metric spaces are obtained in [23, 2].

In $[14,15]$, several classical fixed point theorems and common fixed point theorems have been unified considering a general condition by implicit function.

The study of fixed point for mappings satisfying an implicit relation in G metric spaces is initiated in $[16,17]$ and in other papers.

The notion of $\phi$ - maps is introduced in [9]. In [3], Altun and Turkoglu introduced a new class of implicit relation satisfying a $\phi$ - map.

A general fixed point theorem for mappings satisfying $\phi$ - implicit relations in G - metric spaces is obtained in [18].

The purpose of this paper is to introduce a new type of $\phi$ - implicit relation in $S$ - metric spaces and to prove a general fixed point theorem for a pair of weakly compatible mappings in $S$ - metric spaces, generalizing Theorems 1, 2, 4 [23], Theorems 1-7 [13], Corollary 2.19 [13], Theorems 2.2, 2.4 [19], Theorems $3.2,3.3,3.4[20]$ and other known results.

## 2 Preliminaries

Definition 1 ([21, 22]) A S - metric on a nonempty set X is a function $S: X^{3} \rightarrow \mathbb{R}_{+}$such that for all $x, y, z, a \in X$ :
$\left(S_{1}\right): S(x, y, z)=0$ if and only if $x=y=z$;
$\left(S_{2}\right): S(x, y, z) \leq S(x, x, a)+S(y, y, a)+S(z, z, a)$.
The pair $(\mathrm{X}, \mathrm{S})$ is called a S - metric space.

Example 1 Let $\mathrm{X}=\mathbb{R}$ and $\mathrm{S}(\mathrm{x}, \mathrm{y}, \mathrm{z})=|\mathrm{x}-\mathrm{z}|+|\mathrm{y}-\mathrm{z}|$. Then, $\mathrm{S}(\mathrm{x}, \mathrm{y}, \mathrm{z})$ is a $S$ - metric on $\mathbb{R}$ and is named the usual S - metric on X .

Lemma $1([4,5])$ If S is a S - metric on a nonempty set X , then

$$
S(x, x, y)=S(y, y, x) \text { for all } x, y \in X
$$

Definition 2 ([22]) Let (X,S) be a S - metric space. For $\mathrm{r}>0$ and $\mathrm{x} \in \mathrm{X}$ we define the open ball with center x and radius r , denoted $\mathrm{B}_{\mathrm{S}}(\mathrm{x}, \mathrm{r})$, respectively closed ball, denoted $\overline{\mathrm{B}}_{\mathrm{S}}(\mathrm{x}, \mathrm{r})$, the sets:

$$
B_{S}(x, r)=\{y \in X: S(x, x, y)<r\},
$$

respectively,

$$
\overline{\mathrm{B}}_{S}(x, r)=\{y \in X: S(x, x, y) \leq r\}
$$

The topology induced by $S$ - metric on $X$ is the topology determined by the base of all open balls in $X$.

Definition 3 ([22]) a) A sequence $\left\{x_{n}\right\}$ in a $S$ - metric space $(X, S)$ is convergent to x , denoted $\mathrm{x}_{\mathrm{n}} \rightarrow \mathrm{x}$ or $\lim _{\mathrm{n} \rightarrow \infty} \mathrm{x}_{\mathrm{n}}=\mathrm{x}$, if $\mathrm{S}\left(\mathrm{x}_{\mathrm{n}}, \mathrm{x}_{\mathrm{n}}, \mathrm{x}\right) \rightarrow 0$ as $\mathrm{n} \rightarrow \infty$, that is, for $\varepsilon>0$, there exists $n_{0} \in \mathbb{N}$ such that for all $n \geq n_{0}$ we have $\mathrm{S}\left(\mathrm{x}_{\mathrm{n}}, \mathrm{x}_{\mathrm{n}}, \mathrm{x}\right)<\varepsilon$.
b) $A$ sequence $\left\{x_{n}\right\}$ in $(X, S)$ is a Cauchy sequence if $S\left(x_{n}, x_{n}, x_{m}\right) \rightarrow 0$ as $\mathrm{n}, \mathrm{m} \rightarrow \infty$, that is, for $\varepsilon>0$, there exists $\mathfrak{n}_{0} \in \mathbb{N}$ such that for all $\mathfrak{m}, \mathrm{n} \geq \mathfrak{n}_{0}$ we have $S\left(x_{n}, x_{n}, x_{m}\right)<\varepsilon$.
c) A S - metric space $(\mathrm{X}, \mathrm{S})$ is complete if every Cauchy sequence is convergent.

Example $2(\mathrm{X}, \mathrm{S})$ by Example 1 is complete.
Lemma 2 ([22]) Let $(X, S)$ be a S - metric space. If $\mathrm{x}_{\mathrm{n}} \rightarrow \mathrm{x}$ and $\mathrm{y}_{\mathrm{n}} \rightarrow \mathrm{y}$, then $S\left(x_{n}, x_{n}, y_{n}\right) \rightarrow S(x, x, y)$.

Lemma 3 ([22]) Let $(X, S)$ be a $S$ - metric space and $x_{n} \rightarrow x$. Then $\lim _{n \rightarrow \infty} x_{n}$ is unique.

Lemma 4 ([4]) Let $(X, S)$ be a $S$ - metric space and $\left\{x_{n}\right\}$ be a sequence in X such that

$$
\lim _{n \rightarrow \infty} S\left(x_{n}, x_{n}, x_{n+1}\right)=0 .
$$

If $\left\{\mathrm{x}_{\mathrm{n}}\right\}$ is not a Cauchy sequence, then there exists an $\varepsilon>0$ and two sequences $\left\{\mathfrak{m}_{k}\right\}$ and $\left\{n_{k}\right\}$ of positive integers with $n_{k}>\mathfrak{m}_{k}>k$ such that

$$
S\left(x_{m_{k}}, x_{m_{k}}, x_{n_{k}}\right) \geq \varepsilon, S\left(x_{m_{k-1}}, x_{m_{k-1}}, x_{n_{k}}\right)<\varepsilon
$$

and
(i) $\lim _{n \rightarrow \infty} S\left(x_{m_{k}}, x_{m_{k}}, x_{n_{k}}\right)=\varepsilon$,
(ii) $\lim _{n \rightarrow \infty} S\left(x_{m_{k}}, x_{m_{k}}, x_{n_{k-1}}\right)=\varepsilon$,
(iii) $\lim _{n \rightarrow \infty} S\left(x_{m_{k-1}}, x_{m_{k-1}}, x_{n_{k}}\right)=\varepsilon$,
(iv) $\lim _{n \rightarrow \infty} S\left(x_{m_{k-1}}, x_{m_{k-1}}, x_{n_{k-1}}\right)=\varepsilon$.

Definition $4([9])$ Let $\Phi$ be the set of all functions such that $\phi:[0, \infty) \rightarrow$ $[0, \infty)$ is a nondecreasing function satisfying $\lim _{n \rightarrow \infty} \phi^{n}(\mathrm{t})=0$ for all $\mathrm{t} \in$ $[0, \infty)$. If $\phi \in \Phi$, then $\phi$ is called $\phi$ - mapping. Furthermore, if $\phi \in \Phi$, then:
(i) $\phi(\mathrm{t})<\mathrm{t}$ for all $\mathrm{t} \in(0, \infty)$,
(ii) $\phi(0)=0$.

The following theorems are recently published in [23].
Theorem 1 (Theorem 1 [23]) Let ( $\mathrm{X}, \mathrm{S}$ ) be a S - metric space. Suppose that the mappings $\mathrm{f}, \mathrm{g}: \mathrm{X} \rightarrow \mathrm{X}$ satisfy

$$
\begin{equation*}
S(f x, f y, g z) \leq \phi(\max \{S(g x, g x, f x), S(g y, g y, f y), S(g z, g z, f z)\}) \tag{1}
\end{equation*}
$$

for all $x, y, z \in X$.
If $f(X) \subset g(X)$ and one of $f(X)$ or $g(X)$ is a complete subspace of $X$, then f and g have a unique point of coincidence.

Moreover, if f and g are weakly compatible, then f and g have a unique common fixed point.

Theorem 2 (Theorem 2 [23]) Let (X, S) be a S - metric space. Suppose that the mappings $\mathrm{f}, \mathrm{g}: \mathrm{X} \rightarrow \mathrm{X}$ satisfy

$$
\begin{equation*}
S(f x, f y, f z) \leq \max \{\phi(S(g x, g x, f x)), \phi(S(g y, g y, f y)), \phi(S(g z, g z, f z))\} \tag{2}
\end{equation*}
$$

for all $x, y, z \in X$.
If $f(X) \subset g(X)$ and one of $f(X)$ or $g(X)$ is a complete subspace of $X$, then f and g have a unique point of coincidence. Moreover, if f and g are weakly compatible, then f and g have a unique common fixed point.

Theorem 3 (Theorem 4 [23]) Let $(X, S)$ be a S - metric space. Suppose that the mappings $\mathrm{f}, \mathrm{g}: \mathrm{X} \rightarrow \mathrm{X}$ satisfy

$$
\begin{equation*}
S(f x, f y, f z) \leq k_{1} \phi(S(g x, g x, f x))+k_{2} \phi(S(g y, g y, f y))+k_{3} \phi(S(g z, g z, f z)) \tag{3}
\end{equation*}
$$

for all $x, y, z \in X, k_{1}+k_{2}+k_{3}<1$.
If $\mathrm{f}(\mathrm{X}) \subset \mathrm{g}(\mathrm{X})$ and one of $\mathrm{f}(\mathrm{X})$ or $\mathrm{g}(\mathrm{X})$ is a complete subspace of X , then f and g have a unique point of coincidence. Moreover, if f and g are weakly compatible, then f and g have a unique common fixed point.

Remark 1 1) Since $\phi(\mathrm{t})$ is nondecreasing, then

$$
\phi\left(\max \left\{\mathrm{t}_{2}, \mathrm{t}_{3}, \mathrm{t}_{4}, \mathrm{t}_{5}, \mathrm{t}_{6}\right\}\right)=\max \left\{\phi\left(\mathrm{t}_{2}\right), \phi\left(\mathrm{t}_{3}\right), \phi\left(\mathrm{t}_{4}\right), \phi\left(\mathrm{t}_{5}\right), \phi\left(\mathrm{t}_{6}\right)\right\} .
$$

Hence, Theorem 2 is Theorem 1.
2) By (3) we obtain

$$
\begin{aligned}
S(f x, f y, f z) \leq & \left(k_{1}+k_{2}+k_{3}\right) \max \{\phi(S(g x, g x, f x)) \\
& \phi(S(g y, g y, f y)), \phi(S(g z, g z, f z))\} \\
= & \left(k_{1}+k_{2}+k_{3}\right) \phi(\max \{S(g x, g x, f x), S(g y, g y, f y) \\
& S(g z, g z, f z)\}) \\
\leq & \phi(\max \{\max \{S(g x, g x, f x), S(g y, g y, f y), S(g z, g z, f z)\}\}) .
\end{aligned}
$$

Hence,

$$
S(f x, f y, f z) \leq \phi(\max \{S(g x, g x, f x), S(g y, g y, f y), S(g z, g z, f z)\})
$$

which is the inequality (1). Hence, Theorem 3 is a particular case of Theorem 1 .
3) In the proof of Theorem 1 is used $\mathrm{x}=\mathrm{y}$. Hence in Theorem 1 we have a new form of inequality (1):

$$
S(f x, f x, f y) \leq \phi(\max \{S(g x, g x, f x), g(f y, g y, f y)\})
$$

## $3 \quad \phi$-implicit relations

Let $\mathcal{F}_{\phi}$ be the set of all lower semi - continuous functions $F: \mathbb{R}_{+}^{6} \rightarrow \mathbb{R}$ such that:
$\left(F_{1}\right): \quad F$ is nonincreasing in variable $t_{6}$,
$\left(F_{2}\right): \quad$ There exists $\phi \in \mathcal{F}_{\phi}$ such that for all $u, v \geq 0, F(u, v, v, u, 0,2 u+v) \leq$ 0 implies $u \leq \phi(v)$;
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)>0, \forall t>0$.
In all the following examples, $\left(F_{1}\right)$ is obviously.
Example $3 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{k} \max \left\{\mathrm{t}_{2}, \mathrm{t}_{3}, \ldots, \mathrm{t}_{6}\right\}$, where $\mathrm{k} \in\left[0, \frac{1}{3}\right)$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u-k(u+2 v) \leq 0$. If $u>v$, then $u(1-3 k) \leq 0$, a contradiction. Hence, $u \leq v$, which implies $u \leq 3 k v$ and $F$ satisfies $\left(F_{2}\right)$ for $\phi(t)=3 k t$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t(1-k)>0, \forall t>0$.
Example $4 F\left(t_{1}, \ldots, t_{6}\right)=t_{1}-k \max \left\{t_{2}, t_{3}, t_{4}, \frac{t_{5}+t_{6}}{3}\right\}$, where $k \in[0,1)$.
$\left(F_{2}\right): \quad$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u-k \max \left\{u, v, \frac{2 u+v}{3}\right\} \leq$ 0 . If $u>v$, then $u(1-k) \leq 0$, a contradiction. Hence, $u \leq v$, which implies $u \leq k v$ and $F$ satisfies $\left(F_{2}\right)$ for $\phi(t)=k t$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t(1-k)>0, \forall t>0$.
Example $5 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{at}_{2}-\mathrm{bt}_{3}-\mathrm{ct}_{4}-\mathrm{dt}_{5}-\mathrm{et}_{6}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d}, \mathrm{e} \geq 0$ and $\mathrm{a}+\mathrm{b}+\mathrm{c}+3 \mathrm{e}<1$ and $\mathrm{a}+\mathrm{d}+\mathrm{e}<1$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u-a v-b v-c u-$ $e(2 u+v) \leq 0$. If $u>v$, then $u[1-(a+b+c+3 e)] \leq 0$, a contradiction. Hence, $u \leq v$, which implies $u \leq(a+b+c+3 e) v$ and $F$ satisfies $\left(F_{2}\right)$ for $\phi(t)=(a+b+c+3 e) t$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t[1-(a+d+e)]>0, \forall t>0$.
Example $6 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}^{2}-\mathrm{t}_{1}\left(a \mathrm{t}_{2}+\mathrm{bt}_{3}+\mathrm{ct}_{4}\right)-\mathrm{dt}_{5} \mathrm{t}_{6}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d} \geq$ $0, a+b+c<1$ and $a+d<1$.
$\left(F_{2}\right): \quad$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u^{2}-u(a v+b v+c u) \leq 0$. If $u>v$, then $u^{2}[1-(a+b+c)] \leq 0$, a contradiction. Hence, $u \leq v$, which implies $u \leq(a+b+c) v$ and $F$ satisfies $\left(F_{2}\right)$ for $\phi(t)=(a+b+c) t$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t^{2}[1-(a+d)]>0, \forall t>0$.
Example $7 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}^{2}-\mathrm{at} \mathrm{t}_{2}^{2}-\frac{\mathrm{b} \mathrm{t}_{5} \mathrm{t}_{6}}{1+\mathrm{t}_{3}^{2}+\mathrm{t}_{4}^{2}}$, where $\mathrm{a}, \mathrm{b} \geq 0$ and $\mathrm{a}+\mathrm{b}<1$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u^{2}-a v^{2} \leq 0$, which implies $u \leq \sqrt{a} v$. Hence, $F$ satisfies $\left(F_{2}\right)$ for $\phi(t)=\sqrt{a} t$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t^{2}[1-(a+b)]>0, \forall t>0$.
In the following examples, if $\phi \in \Phi$, then $F$ satisfy properties $\left(F_{1}\right),\left(F_{2}\right),\left(F_{3}\right)$.
Example $8 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\phi\left(\max \left\{\mathrm{t}_{2}, \mathrm{t}_{3}, \mathrm{t}_{4}, \frac{\mathrm{t}_{5}+\mathrm{t}_{6}}{3}\right\}\right)$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and

$$
F(u, v, v, u, 0,2 u+v)=u-\phi\left(\max \left\{u, v, \frac{2 u+v}{3}\right\}\right) \leq 0 .
$$

If $u>v$, then $u \leq \phi(u)<u$, a contradiction. Hence, $u \leq v$, which implies $u \leq \phi(v)$.
$\left(\mathrm{F}_{3}\right): \quad \mathrm{F}(\mathrm{t}, \mathrm{t}, 0,0, \mathrm{t}, \mathrm{t})=\mathrm{t}-\phi(\mathrm{t})>0, \forall \mathrm{t}>0$.
Example $9 F\left(t_{1}, \ldots, t_{6}\right)=t_{1}-\phi\left(\max \left\{t_{2}, \frac{t_{3}+t_{4}}{2}, \frac{t_{5}+t_{6}}{3}\right\}\right)$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and

$$
\mathrm{F}(\mathrm{u}, v, v, u, 0,2 u+v)=u-\phi\left(\max \left\{u, \frac{u+v}{2}, \frac{2 u+v}{3}\right\}\right) \leq 0
$$

If $u>v$, then $u \leq \phi(u)<u$, a contradiction. Hence, $u \leq v$, which implies $u \leq \phi(v)$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t-\phi(t)>0, \forall t>0$.
Example $10 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\phi\left(\mathrm{at}_{2}+\mathrm{b} \max \left\{\mathrm{t}_{3}, \mathrm{t}_{4}\right\}+\mathrm{c} \max \left\{\mathrm{t}_{5}, \mathrm{t}_{6}\right\}\right)$, where $\mathrm{a}, \mathrm{b}, \mathrm{c} \geq 0$ and $\mathrm{a}+\mathrm{b}+3 \mathrm{c}<1$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and

$$
F(u, v, v, u, 0,2 u+v)=u-\phi(a v+b \max \{u, v\}+c(2 u+v)) \leq 0 .
$$

If $u>v$, then $u-\phi((a+b+3 c) u) \leq 0$, which implies $u \leq \phi(u)<u$, a contradiction. Hence, $u \leq v$ and $u \leq \phi(v)$.
$\left(F_{3}\right): F(t, t, 0,0, t, t)=t-\phi(a t+c t) \geq t-\phi((a+b+3 c) t) \geq t-$ $\phi(\mathrm{t})>0, \forall \mathrm{t}>0$.

Example $11 F\left(t_{1}, \ldots, t_{6}\right)=t_{1}-\phi\left(a \sqrt{t_{1} t_{2}}+b \sqrt{t_{3} t_{4}}+c \sqrt{t_{5} t_{6}}\right)$, where $a, b, c$ $\geq 0$ and $\mathrm{a}+\mathrm{b}+\mathrm{c}<1$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u-\phi(a \sqrt{u v}+b \sqrt{u v}) \leq 0$. If $u>v$, then $u \leq \phi((a+b) u)<u$, a contradiction. Hence, $u \leq v$, which implies $u \leq \phi(v)$.
$\left(F_{3}\right): F(t, t, 0,0, t, t)=t-\phi((a+c) t) \geq t-\phi((a+b+c) t) \geq t-$ $\phi(\mathrm{t})>0, \forall \mathrm{t}>0$.

Example $12 F\left(t_{1}, \ldots, t_{6}\right)=t_{1}-\phi\left(a t_{2}, \frac{b \sqrt{t_{5} t_{6}}}{1+\mathrm{t}_{3}+\mathrm{t}_{4}}\right)$, where $\mathrm{a}, \mathrm{b} \geq 0$ and $a+b<1$.
$\left(F_{2}\right):$ Let $u, v \geq 0$ and $F(u, v, v, u, 0,2 u+v)=u-\phi(a v) \leq 0$. If $u>v$, then $u-\phi(a v) \leq 0$ implies $u \leq \phi(u)<u$, a contradiction. Hence, $u \leq v$, which implies $u \leq \phi(v)$.
$\left(F_{3}\right): \quad F(t, t, 0,0, t, t)=t-\phi((a+b) t) \geq t-\phi(t)>0, \forall t>0$.
In the following examples, the proofs are similar to the proof of Example 12 and thus are omitted.

Example $13 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{at}_{2}-\mathrm{b} \max \left\{\mathrm{t}_{3}, \mathrm{t}_{4}, \mathrm{t}_{5}, \mathrm{t}_{6}\right\}$, where $\mathrm{a}, \mathrm{b} \geq 0$ and $a+3 b<1$.

If $\mathrm{F}(\mathrm{u}, v, v, u, 0,2 u+v) \leq 0$, then we have $u \leq \phi(v)$, where $\phi(\mathrm{t})=(\mathrm{a}+3 \mathrm{~b}) \mathrm{t}$.
Example $14 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{at}_{2}-\mathrm{bt}_{3}-\mathrm{ct}_{4}-\mathrm{d} \max \left\{\mathrm{t}_{5}, \mathrm{t}_{6}\right\}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d} \geq$ 0 and $\mathrm{a}+\mathrm{b}+\mathrm{c}+3 \mathrm{~d}<1$.

If $\mathrm{F}(\mathrm{u}, v, v, \mathfrak{u}, 0,2 \mathfrak{u}+v) \leq 0$ then we have $u \leq \phi(v)$, where $\phi(\mathrm{t})=(\mathrm{a}+\mathrm{b}+$ $\mathrm{c}+3 \mathrm{~d}) \mathrm{t}$.

Example $15 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{at}_{2}-\mathrm{d} \max \left\{\mathrm{t}_{3}, \mathrm{t}_{4}\right\}-\mathrm{bt}_{5}-\mathrm{ct}_{6}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d} \geq$ $0, a+3 c+d \geq 0, a+3 c+d<1$ and $a+b+c<1$.

If $F(u, v, v, u, 0,2 u+v) \leq 0$ then $u \leq \phi(v)$, where $\phi(t)=(a+3 c+d) t$.
Example $16 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{at}_{2}-\mathrm{bt}_{3}-\mathrm{et}_{4}-\mathrm{ct}_{5}-\mathrm{dt}_{6}-\mathrm{f} \max \left\{\mathrm{t}_{2}, \mathrm{t}_{3}, \ldots, \mathrm{t}_{6}\right\}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d}, \mathrm{e}, \mathrm{f} \geq 0, \mathrm{a}+\mathrm{b}+\mathrm{e}+3 \mathrm{~d}+3 \mathrm{f}<1$ and $\mathrm{a}+\mathrm{c}+\mathrm{e}+\mathrm{f}<1$.

If $\mathrm{F}(\mathrm{u}, v, v, \mathfrak{u}, 0,2 \mathfrak{u}+v) \leq 0$ then $\mathfrak{u} \leq \phi(v)$, where $\phi(\mathrm{t})=(\mathrm{a}+\mathrm{b}+\mathrm{e}+3 \mathrm{~d}$ $+3 f) \mathrm{t}$.

Example $17 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{a}\left(\mathrm{t}_{5}+\mathrm{t}_{6}\right)-\mathrm{b} \mathrm{t}_{2}-\mathrm{c} \max \left\{\mathrm{t}_{3}, \mathrm{t}_{4}\right\}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c} \geq$ 0 and $3 \mathrm{a}+\mathrm{b}+\mathrm{c}<1$.

If $\mathrm{F}(\mathrm{u}, v, v, \mathfrak{u}, 0,2 \mathfrak{u}+v) \leq 0$ then $\mathfrak{u} \leq \phi(v)$, where $\phi(\mathrm{t})=(3 \mathrm{a}+\mathrm{b}+\mathrm{c}) \mathrm{t}$.
Example $18 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{a}\left(\mathrm{t}_{3}+\mathrm{t}_{4}\right)-\mathrm{b} \mathrm{t}_{2}-\mathrm{c} \max \left\{\mathrm{t}_{5}, \mathrm{t}_{6}\right\}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c} \geq$ 0 and $2 \mathrm{a}+\mathrm{b}+3 \mathrm{c}<1$.

If $\mathrm{F}(\mathrm{u}, v, v, \mathfrak{u}, 0,2 u+v) \leq 0$ then $u \leq \phi(v)$, where $\phi(\mathrm{t})=(2 \mathrm{a}+\mathrm{b}+3 \mathrm{c}) \mathrm{t}$.
Example $19 \mathrm{~F}\left(\mathrm{t}_{1}, \ldots, \mathrm{t}_{6}\right)=\mathrm{t}_{1}-\mathrm{a} \max \left\{\mathrm{t}_{4}+\mathrm{t}_{5}, \mathrm{t}_{3}+\mathrm{t}_{6}\right\}-\mathrm{bt}_{2}$, where $\mathrm{a}, \mathrm{b}, \mathrm{c} \geq 0$ and $4 \mathrm{a}+\mathrm{b}<1$.

If $\mathrm{F}(\mathrm{u}, v, v, \mathfrak{u}, 0,2 \mathfrak{u}+v) \leq 0$ then $u \leq \phi(v)$, where $\phi(\mathrm{t})=(4 \mathrm{a}+\mathrm{b}) \mathrm{t}$.

## 4 Main results

Lemma 5 ([1]) Let f and g be weakly compatible self mappings of a nonempty set X . If f and g have a unique point of coincidence $w=\mathrm{fx}=\mathrm{gx}$ for some $x \in X$, then $w$ is the unique common fixed point of f and g .

Theorem 4 Let $(\mathrm{X}, \mathrm{S})$ be a S - metric space and $\mathrm{f}, \mathrm{g}: \mathrm{X} \rightarrow \mathrm{X}$ such that

$$
\begin{equation*}
F\binom{S(f x, f x, f y), S(g x, g x, g y), S(g x, g x, f x),}{S(g y, g y, f y), S(g y, g y, f x), S(g x, g x, f y)} \leq 0 \tag{4}
\end{equation*}
$$

for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ and some $\mathrm{F} \in \mathcal{F}_{\phi}$.
If $f(\mathrm{X}) \subset \mathrm{g}(\mathrm{X})($ or $\mathrm{g}(\mathrm{X}) \subset \mathrm{f}(\mathrm{X}))$ and $\mathrm{g}(\mathrm{X})$ (or $\mathrm{f}(\mathrm{X})$ ) is a complete subspace of $(\mathrm{X}, \mathrm{S})$, then f and g have a unique point of coincidence. Moreover, if f and g are weakly compatible, then f and g have a unique common fixed point.

Proof. Let $x_{0}$ be an arbitrary point of $X$. Since $f(X) \subset g(X)$, there exists $\mathrm{x}_{1} \in \mathrm{X}$ such that $\mathrm{f} \mathrm{x}_{0}=\mathrm{g} \mathrm{x}_{1}$. Continuing this process we define the sequence $\left\{x_{n}\right\}$ satisfying

$$
f x_{n}=g x_{n+1} \text { for } n \in \mathbb{N} .
$$

Then, by (4) for $x=x_{n-1}$ and $y=x_{n}$ we have

$$
\begin{align*}
& F\binom{S\left(f x_{n-1}, f x_{n-1}, f x_{n}\right), S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right), S\left(g x_{n-1}, g x_{n-1}, f x_{n-1}\right),}{S\left(g x_{n}, g x_{n}, f x_{n}\right), S\left(g x_{n}, g x_{n}, f x_{n-1}\right), S\left(g x_{n-1}, g x_{n-1}, f x_{n}\right)} \leq 0 \\
& F\binom{S\left(g x_{n}, g x_{n}, g x_{n+1}\right), S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right), S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right),}{S\left(g x_{n}, g x_{n}, g x_{n+1}\right), 0, S\left(g x_{n-1}, g x_{n-1}, g x_{n+1}\right)} \leq 0 \tag{5}
\end{align*}
$$

By $\left(S_{2}\right)$ and Lemma 1 we have

$$
S\left(g x_{n-1}, g x_{n-1}, g x_{n+1}\right) \leq 2 S\left(g x_{n}, g x_{n}, g x_{n+1}\right)+S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right)
$$

By (5) and $\left(F_{1}\right)$ we obtain
$F\binom{S\left(g x_{n}, g x_{n}, g x_{n+1}\right), S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right), S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right)}{,S\left(g x_{n}, g x_{n}, g x_{n+1}\right), 0,2 S\left(g x_{n}, g x_{n}, g x_{n+1}\right)+S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right)} \leq 0$.
By $\left(F_{2}\right)$ we obtain

$$
S\left(g x_{n}, g x_{n}, g x_{n+1}\right) \leq \phi\left(S\left(g x_{n-1}, g x_{n-1}, g x_{n}\right)\right), \text { for } n=1,2, \ldots
$$

which implies

$$
S\left(g x_{n}, g x_{n}, g x_{n+1}\right) \leq \phi^{n}\left(S\left(g x_{0}, g x_{0}, g x_{1}\right)\right)
$$

Letting n tend to infinity we obtain

$$
\lim _{n \rightarrow \infty} S\left(g x_{n}, g x_{n}, g x_{n+1}\right)=0
$$

We prove that $\left\{g x_{n}\right\}$ is a Cauchy sequence in $g(X)$. Suppose that $\left\{g x_{n}\right\}$ is not a Cauchy sequence. Then, by Lemma 4, there exists an $\varepsilon>0$ and two sequences $m_{k}$ and $n_{k}$ with $n_{k}>m_{k}>k$ and $S\left(x_{m_{k}}, x_{m_{k}}, x_{n_{k}}\right) \geq \varepsilon$ and $S\left(x_{m_{k}-1}, x_{m_{k}-1}, x_{\mathfrak{n}_{k}}\right)<\varepsilon$ and satisfying the inequalities (i) - (iv) by Lemma 4.

By (4) for $x=x_{m_{k}-1}$ and $y=x_{n_{k}-1}$ we have

$$
\begin{gather*}
F\left(\begin{array}{c}
S\left(f x_{m_{k}-1}, f x_{m_{k}-1}, f x_{n_{k}-1}\right), S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, g x_{n_{k}-1}\right), \\
S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, f x_{m_{k}-1}\right), S\left(g x_{n_{k}-1}, g x_{n_{k}-1}, f x_{n_{k}-1}\right), \\
S\left(g x_{n_{k}-1}, g x_{n_{k}-1}, f x_{m_{k}-1}\right), S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, f x_{n_{k}-1}\right)
\end{array}\right) \leq 0 \\
F\left(\begin{array}{c}
S\left(g x_{m_{k}}, g x_{m_{k}}, g x_{n_{k}}\right), S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, g x_{n_{k}-1}\right), \\
S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, g x_{m_{k}}\right), S\left(g x_{n_{k}-1}, g x_{n_{k}-1}, g x_{n_{k}}\right), \\
S\left(g x_{n_{k}-1}, g x_{n_{k}-1}, g x_{m_{k}}\right), S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, g x_{n_{k}}\right)
\end{array}\right) \leq 0 . \tag{6}
\end{gather*}
$$

By Lemma 1,

$$
S\left(g x_{m_{k}-1}, g x_{m_{k}-1}, g x_{n_{k}}\right)=S\left(g x_{n_{k}}, g x_{n_{k}}, g x_{m_{k}-1}\right)
$$

and

$$
S\left(g x_{n_{k}-1}, g x_{n_{k}-1}, g x_{m_{k}}\right)=S\left(g x_{m_{k}}, g x_{m_{k}}, g x_{n_{k}-1}\right) .
$$

Letting n tend to infinity in (6) we obtain

$$
F(\varepsilon, \varepsilon, 0,0, \varepsilon, \varepsilon) \leq 0
$$

a contradiction of $\left(F_{3}\right)$.
Hence, $\left\{g x_{n}\right\}$ is a Cauchy sequence in $g(X)$. Since $g(X)$ is complete, then $\left\{g x_{n}\right\}$ is convergent to a point $t \in g(X)$. Hence, there exists $p \in X$ such that $\mathrm{gp}=\mathrm{t}$ and $\lim _{\mathrm{n} \rightarrow \infty} \mathrm{g} \mathrm{x}_{\mathrm{n}}=\mathrm{gp}$. We prove that $\mathrm{fp}=\mathrm{gp}$.

By (4) for $x=x_{n}$ and $y=p$ we have

$$
F\binom{S\left(g x_{n}, g x_{n}, f p\right), S\left(g x_{n}, g x_{n}, g p\right), S\left(g x_{n}, g x_{n}, f x_{n}\right),}{S(g p, g p, f p), S\left(g p, g p, f x_{n}\right), S\left(g x_{n}, g x_{n}, f p\right)} \leq 0
$$

Letting $\mathfrak{n}$ tend to infinity we obtain

$$
F(S(g p, g p, f p), 0,0, S(g p, g p, f p), 0, S(g p, g p, f p)) \leq 0
$$

By $\left(F_{1}\right)$ we have

$$
F(S(g p, g p, f p), 0,0, S(g p, g p, f p), 0,2 S(g p, g p, f p)) \leq 0
$$

which implies $S(g p, g p, f p)=0$. Hence $g p=f p=t$.
We prove that $t$ is the unique point of coincidence of $f$ and $g$. Suppose that there exists $z=f w=g w$. By (4) we obtain

$$
\begin{gathered}
F\binom{S(f p, f p, f w), S(g p, g p, g w), S(g p, g p, f p),}{S(g w, g w, f w), S(g w, g w, f p), S(g p, g p, f w)} \leq 0 \\
F(S(t, t, z), S(t, t, z), 0,0, S(z, z, t), S(t, t, z)) \leq 0
\end{gathered}
$$

By Lemma 1 we have

$$
F(S(t, t, z), S(t, t, z), 0,0, S(t, t, z), S(t, t, z)) \leq 0
$$

a contradiction of $\left(F_{3}\right)$ if $S(t, t, z)>0$. Hence, $z=t$ and $t$ is the unique point of coincidence of $f$ and $g$.

Moreover, if f and g are weakly compatible, then by Lemma $5, \mathrm{f}$ and g have a unique common fixed point $t$.

If $\phi(t)=k t, k \in[0,1)$, by Example 8 and Theorem 4 we obtain
Corollary 1 Let $(\mathrm{X}, \mathrm{S})$ be a S - metric space and $\mathrm{f}, \mathrm{g}: \mathrm{X} \rightarrow \mathrm{X}$ such that

$$
S(f x, f x, f y) \leq k \max \left\{\begin{array}{c}
S(g x, g x, g y), S(g x, g x, f x), S(g y, g y, f y)  \tag{7}\\
\frac{S(g y, g y, f x)+S(g x, g x, f y)}{3}
\end{array}\right\}
$$

where $\mathrm{k} \in[0,1)$.
If $f(X) \subset g(X)($ or $g(X) \subset f(X))$ and $g(X)(o r f(X))$ is a complete subspace of $(\mathrm{X}, \mathrm{S})$, then f and g have a unique point of coincidence. Moreover, if f and g are weakly compatible, then f and g have a unique common fixed point.

Example 20 Let $\mathrm{X}=\mathbb{R}$ and $\mathrm{S}(\mathrm{x}, \mathrm{y}, \mathrm{z})=|\mathrm{x}-\mathrm{z}|+|\mathrm{y}-\mathrm{z}|$. Then $\mathrm{S}(\mathrm{X})$ is a complete $S$ - metric space. Let $\mathrm{fx}=2 \mathrm{x}-2$, $\mathrm{gx}=3 \mathrm{x}-4$. Then $\mathrm{f}(\mathrm{X})=\mathbb{R}$, $\mathrm{g}(\mathrm{X})=\mathbb{R}$ and $\mathrm{f}(\mathrm{X}) \subset \mathrm{g}(\mathrm{X})$. If $\mathrm{f} \mathrm{x}=\mathrm{gx}$, then $\mathrm{x}=2$ which implies $\mathcal{C}(\mathrm{f}, \mathrm{g})=\{2\}$ and $\mathrm{fg} 2=\mathrm{gf} 2=2$ and $\mathrm{x}=2$ is the unique point of coincidence of f and g and f and g are weakly compatible. On the other hand, $\mathrm{S}(\mathrm{fx}, \mathrm{fx}, \mathrm{fy})=4|\mathrm{x}-\mathrm{z}|$ and $S(g x, g x, g y)=6|x-y|$. Hence, $S(f x, f x, f y) \leq k S(g x, g x, g y)$, for $k \in$ $\left[\frac{2}{3}, 1\right)$. This implies

$$
S(f x, f x, f y) \leq k \max \left\{\begin{array}{c}
S(g x, g x, g y), S(g x, g x, f x), S(g y, g y, f y), \\
\frac{S(g y, g y, f x)+S(g x, g x, f y)}{3}
\end{array}\right\}
$$

for $\mathrm{k} \in\left[\frac{2}{3}, 1\right.$. By Corollary 1, f and g have a unique common fixed point $x=2$.

If $g(x)=x$, then by Theorem 4 we obtain
Theorem 5 Let $(X, S)$ be a complete S - metric space and $\mathrm{f}: \mathrm{X} \rightarrow \mathrm{X}$ such that
$F(S(f x, f x, f y), S(x, x, y), S(x, x, f x), S(y, y, f y), S(y, y, f x), S(x, x, f y)) \leq 0$, for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ and some $\mathrm{F} \in \mathcal{F}_{\phi}$.

Then f has a unique fixed point.
Corollary 2 Let ( $\mathrm{X}, \mathrm{S}$ ) be a complete S - metric space and $\mathrm{f}: \mathrm{X} \rightarrow \mathrm{X}$ such that
$S(f x, f x, f y) \leq k \max \{S(x, x, y), S(x, x, f x), S(y, y, f y), S(x, x, f y), S(x, x, f y)\}$, for all $\mathrm{x}, \mathrm{y} \in \mathrm{X}$ and $\mathrm{k} \in\left[0, \frac{1}{3}\right]$. Then f has a unique fixed point.

Proof. The proof follows by Theorem 5 and Example 4.

Remark 2 1) By Examples 13 - 19 and Theorem 4 we obtain Theorems 1-7 [13].
2) By Example 4 and Theorem 4 we obtain Corollary 2.19 [13].
3) By Example 5 and Theorem 4 we obtain Theorems 2.2, 2.4 [19] and Theorems 3.2, 3.3, 3.4 [20].

## References

[1] M. Abbas and B. E. Rhoades, Common fixed point results for noncommuting mappings without continuity in generalized metric spaces, Appl. Math. Comput., 215 (5) (2009), 262-269.
[2] J. M. Afra, Fixed point theorems in S - metric spaces, Theory Approx. Appl., 10 (1) (2014), 57-68.
[3] I. Altun and D. Turkoglu, Some fixed point theorems for weakly compatible mappings satisfying an implicit relation, Taiwanesse J. Math., 13 (4) (2009), 1291-1304.
[4] G. V. R. Babu and B. K. Leta, Fixed points of $(\alpha, \psi, \varphi)$ - generalized weakly contractive maps and property (P) in $S$ - metric spaces, Filomat, 31 (14) (2017), 4469-4481.
[5] B. C. Dhage, Generalized metric space and mappings with fixed point, Bull. Calcutta Math. Soc., 84 (1992), 329-336.
[6] B. C. Dhage, Generalized metric space and topological structures I, An. Ştiinţ. Univ. Al. I. Cuza Iaşi, Mat., 46 (1) (2000), 3-20.
[7] N. V. Dung, N. T. Hieu, S. Radojević, Fixed point theorems for g monotone maps on partially ordered S - metric spaces, Filomat, 28 (9) (2014), 1885-1898.
[8] G. Jungck, Compatible mappings and common fixed points, Int. J. Math. Math. Sci., 9 (1986), 771-779.
[9] J. Matkowski, Fixed point theorems for mappings with a contractive iterate at a point, Proc. Amer. Math. Soc., 62 (2) (1997), 344-348.
[10] Z. Mustafa, Common fixed point of weakly compatible mappings in G metric spaces, Appl. Math. Sci., 92, 6 (2006), 4589-4601.
[11] Z. Mustafa and B. Sims, Some remarks concerning D - metric spaces, Proc. Conf. Fixed Point Theory Appl., Valencia (Spain), 2013, 189-198.
[12] Z. Mustafa and B. Sims, A new approach to generalized metric spaces, $J$. Nonlinear Convex Anal., 7 (2006), 280-297.
[13] N. Y. Özgür and N. Taş, Some generalizations on fixed point theorems on S - metric spaces, In: Rassias T., Pardalos P. (eds) Essays in Mathematics and its Applications. Springer, Cham, 229-261.
[14] V. Popa, Fixed point theorems for implicit contractive mappings, Stud. Cerc. Ştiinţ. Ser. Mat., Univ. Bacău, 7 (1997), 129-133.
[15] V. Popa, Some fixed point theorems for compatible mappings satisfying an implicit relation, Demonstr. Math., 32 (1) (1999), 157-163.
[16] V. Popa, A general fixed point theorem for several mappings in G - metric spaces, Sci. Stud. Res. Ser. Math. Inform., 32 (1) (2011), 205-214.
[17] V. Popa and A.-M. Patriciu, A general fixed point theorem for pairs of weakly compatible mappings in G-metric spaces, J. Nonlinear Sci. Appl., 5 (2) (2012), 151-160.
[18] V. Popa and A.-M. Patriciu, A general fixed point theorem for mappings satisfying an $\phi$ - implicit relation in G-metric spaces, Gazi Univ. J. Sci., 25 (2012), 401-408.
[19] K. Prudhvi, Fixed point results in S - metric spaces, Univer. J. Comput. Math., 3 (2015), 19-21.
[20] K. Prudhvi, Some fixed point results in S - metric spaces, J. Math. Sci. Appl., 4 (1) (2016), 1-3.
[21] S. Sedghi and N. V. Dung, Fixed point theorems on S - metric spaces, Mat. Vesn., 66 (1) (2014), 113-124.
[22] S. Sedghi, N. Shobe and A. Aliouche, A generalization of fixed point theorems in S - metric spaces, Mat. Vesn., 64 (3) (2012), 258-266.
[23] S. Sedghi, M. M. Rezaee, T. Došenović, S. Radenović, Common fixed point theorems for contractive mappings satisfying $\phi-$ maps in $S$ - metric spaces, Acta Univ. Sapientiae, Math., 8 (2) (2016), 298-311.

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# On Euler products with smaller than one exponents 

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#### Abstract

Investigation has been made regarding the properties of the $\prod_{p \leq n}\left(1 \pm 1 / p^{s}\right)$ products over the prime numbers, where we fix the $s \in \overline{\mathbb{R}}$ exponent, and let the $n \geq 2$ natural bound grow toward positive infinity. The nature of these products for the $s \geq 1$ case is known. We get approximations for the case when $s \in[1 / 2,1)$, furthermore different observations for the case when $s<1 / 2$.


## 1 Introduction

In this article, we will investigate the asymptotical properties of Euler products. More precisely, we are going to look at how does the

$$
\begin{equation*}
\prod_{p \leq n}\left(1 \pm \frac{1}{p^{s}}\right) \tag{1}
\end{equation*}
$$

products over the prime numbers behave asymptotically, when we fix the $s \in \mathbb{R}$ exponent, and let the $n \geq 2$ natural bound grow toward positive infinity.

Due to the connection with the Riemann zeta function and Dirichlet series, the "classical" Euler products were and are the subject of a thorough investigation. Some of the results concerning them can describe the nature of the products which we will examine, so we are going to shortly sum up the properties which can be already stated based on these results.

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First we concentrate on the negative case of expression (1). It is well known that

$$
\lim _{n \rightarrow \infty} \prod_{p \leq n}\left(1-\frac{1}{p^{s}}\right)=\frac{1}{\zeta(s)}
$$

when $s>1$, see [2] section 11.5. (This result dates back to Euler, who pointed out this connection in [9] by using simple algebraic manipulations on - what is now known as - the Dirichlet series form of the Riemann zeta function.) As when $s=1$, Mertens proved his infamous result in [12], stating that

$$
\lim _{n \rightarrow \infty} \ln n \prod_{p \leq n}\left(1-\frac{1}{p}\right)=e^{-\gamma}
$$

holds, which is usually referred to as Mertens' third theorem. (To obtain this result, one has to know how the $\zeta(s)$ Riemann zeta function behaves in the neighbourhood of its pole at $s=1$.) When $s=0$, then the product is zero, and when $s$ is negative, then the product diverges.

Concerning the positive case of expression (1), one can use the

$$
\begin{equation*}
\left(1+\frac{1}{p^{s}}\right)\left(1-\frac{1}{p^{s}}\right)=1-\frac{1}{p^{2 s}} \tag{2}
\end{equation*}
$$

equation when $s \in \mathbb{R}$, to transform the results from the negative case. Because of this, we get that

$$
\lim _{n \rightarrow \infty} \prod_{p \leq n}\left(1+\frac{1}{p^{s}}\right)=\frac{\zeta(s)}{\zeta(2 s)}
$$

when $s>1$, and also that

$$
\lim _{n \rightarrow \infty} \frac{1}{\ln n} \prod_{p \leq n}\left(1+\frac{1}{p}\right)=\frac{6}{\pi^{2}} e^{\gamma}
$$

holds when $s=1$. (These results are not new, see again section 11.5 in [2].) When $s$ is zero, the product grows as $2^{\pi(n)}$, where $\pi$ is the prime counting function.

What remains in both cases is when $s \in(0,1)$, and $s<0$ in the positive case of expression (1). We are going to concentrate on the positive case, because equation (2) can be applied to transform our results back to the negative case. We will rely on the following theorem, see theorem 2.7.1 from [3].

Theorem 1 Let f be continuously differentiable on an open interval containing $[2, \infty)$, and let $\pi(x)=\operatorname{Li}(x)+\epsilon(x)$, where $\operatorname{Li}(x)=\int_{2}^{x} d t / \ln (t)$ is the offset logarithmic integral. Now if $x \geq 2$, then

$$
\sum_{p \leq x} f(p)=\int_{2}^{x} \frac{f(t)}{\ln (t)} d t+\epsilon(x) f(x)-\int_{2}^{x} \epsilon(t) f^{\prime}(t) d t
$$

The precision which we can achieve while applying this theorem depends heavily on the applied $\epsilon$ error term. Most of the results in this area give absolute bound to the $\pi(x)-\operatorname{li}(x)=\varepsilon(x)$ error term, where

$$
\operatorname{li}(x)=\lim _{h \rightarrow 0^{+}}\left(\int_{0}^{1-h} \frac{d t}{\ln t}+\int_{1+h}^{x} \frac{d t}{\ln t}\right)
$$

is the logarithmic integral. Take note that the applied $\epsilon$ and the later $\varepsilon$ differs; we are going to use the later $\varepsilon$ throughout the paper. Because of this, we have to substitute the $\epsilon$ error term in theorem 1 as

$$
\epsilon(x)=\pi(x)-\operatorname{Li}(x)=\pi(x)-\operatorname{li}(x)+\operatorname{li}(2)=\varepsilon(x)+\operatorname{li}(2)
$$

later on. Take note that the value of $\mathrm{li}(2)$ is 1.04516378 approximately. Regarding the $\varepsilon$ error term, Koch showed in [10], that if the Riemann hypothesis is true, then $\varepsilon(x) \in \mathcal{O}(\sqrt{x} \ln x)$. This has been made more precise by Schoenfeld in [14], showing that

$$
\begin{equation*}
|\varepsilon(x)|<\frac{\sqrt{x} \ln x}{8 \pi} \tag{3}
\end{equation*}
$$

holds for all $x \geq 2657$. As of now, this is the best possible error bound depending on the validity of the Riemann hypothesis. A weaker result from the same article of Koch states that

$$
|\varepsilon(x)|<\mathcal{O}\left(x^{1 / 2+\sigma}\right)
$$

for all $\sigma>0$, if the Riemann hypothesis is true. Kotnik in his [11] article improves this by conjecturing that even

$$
\begin{equation*}
|\varepsilon(x)|<\sqrt{x} \tag{4}
\end{equation*}
$$

holds for all $x \geq 2$ according to his investigations. We will use this later inequality in our calculations. Now we give our results for the $s \in[1 / 2,1)$ case.

Proposition 1 If the conjecture of Kotnik is true, then

$$
\begin{equation*}
\lim _{n \rightarrow \infty} e^{-\operatorname{li}\left(n^{1-s}\right)} \prod_{p \leq n}\left(1+\frac{1}{p^{s}}\right)=\varphi(s) e^{\mathcal{O}\left(\Gamma\left(1-\frac{1}{2 s}\right)\right)} \tag{5}
\end{equation*}
$$

holds when $\mathrm{s} \in(1 / 2,1)$, where

$$
\begin{equation*}
\varphi(s)=\frac{\sqrt{2 s-1}}{(1-s) \sqrt{\ln 2}} \tag{6}
\end{equation*}
$$

furthermore

$$
\begin{equation*}
\prod_{p \leq n}\left(1+\frac{1}{\sqrt{\mathfrak{p}}}\right)=e^{\mathrm{li}(\sqrt{n})+\mathcal{O}(\ln n)} \tag{7}
\end{equation*}
$$

holds when $\mathrm{s}=1 / 2$.
In the case when Kotnik's conjecture would turn out to be false, one could fall back to using the result of Schoenfeld, see inequality (3), which would yield similar results, but with a more complex right hand side in equation (5), furthermore a much rougher asymptotic in place of equation (7). The plot of the $\varphi(s)$ function for $s \in(1 / 2,1)$ can be seen on figure 1 .

Take note that the gamma function on the right hand side of equation (5) goes to positive infinity as $s$ approaches $1 / 2$ from the right, and $\varphi(s)$ goes to positive infinity as $s$ approaches 1 from the left. When $s$ is not near $1 / 2$ or 1 , the right hand side of equation (5) is smaller than a constant depending on $s$, because next to the $\Gamma$ function, we only have constant terms hidden behind the asymptotic, see the proof in section 2.

It is noted that to obtain the third theorem of Mertens, one has to know how the Riemann zeta function behaves in the neighbourhood of its pole. Here, the Riemann hypothesis is a much stronger assumption, which relies on the exact behaviour of the Riemann zeta function in the critical strip. (For more information about the critical strip, see for example [7].)

As for the cases when $s \in(0,1 / 2)$ and when $s<0$, we are going to get much rougher results. These will be more like observations, and we are going to give them in section 3 , where we will state our remarks.


Figure 1: Plot of the $\varphi(s)$ function for $s \in(1 / 2,1)$ from equation (6). Take note that $\varphi(s)$ approaches positive infinity as $s$ approaches 1 from the left.

Our results have strong connection with the logarithmic integral, which is not so surprising due to the fact that we apply theorem 1 to obtain them. We know, see either chapter 5, equation 5.1.3 and 5.1.10 in [1], or article [6], that

$$
\begin{equation*}
\operatorname{li}(x)=\gamma+\ln \ln x+\sum_{k=1}^{\infty} \frac{\ln ^{k} x}{k!k} \tag{8}
\end{equation*}
$$

holds when $x>1$. As one approaches $x=1$ either from the left or the right, the $\operatorname{li}(x)$ logarithmic integral grows toward negative infinity. Using similar arguments as in [6], one can derive a formula, which is very similar to equation (8), in the case when $x \in(0,1)$.

Lemma 1 When $x \in(0,1)$, then

$$
\operatorname{li}(x)=\gamma+\ln \ln \frac{1}{x}+\sum_{k=1}^{\infty} \frac{\ln ^{k} x}{k!k}
$$

holds.
Proof. Let $x \in(0,1)$. By substituting $t$ with $e^{-u}$, we get that

$$
\operatorname{li}(x)=\int_{0}^{x} \frac{1}{\ln t} d t=-\int_{-\ln x}^{\infty} \frac{e^{-u}}{u} d u=-\int_{0}^{\infty} \frac{e^{-u}}{u} d u+\int_{0}^{-\ln x} \frac{e^{-u}}{u} d u
$$

equalities hold, because $-\ln x>0$ as $x \in(0,1)$. By splitting the first integral on the right hand side, we get

$$
\begin{equation*}
-\int_{0}^{1} \frac{e^{-u}}{u} d u-\int_{1}^{\infty} \frac{e^{-u}}{u} d u+\int_{0}^{-\ln x} \frac{e^{-u}}{u} d u+\int_{\ln x}^{-\ln x} \frac{1}{u} d u \tag{9}
\end{equation*}
$$

where one should observe that the last term which we have added is zero. We want to introduce $\gamma$ into this expression, and because

$$
\gamma=\int_{0}^{1} \frac{1-e^{-u}}{u} d u-\int_{1}^{\infty} \frac{e^{-u}}{u} d u
$$

holds, see page 103 of [5], we want to "cut down" $\int_{0}^{1} 1 / u d u$ from the last term in expression (9). Two scenarios can occur based on the value of $x$. If $x \in\left(0, e^{-1}\right)$, then $-\ln x \geq 1$, so one can do the

$$
\int_{\ln x}^{-\ln x} \frac{1}{\mathfrak{u}} d \mathfrak{u}=\int_{\ln x}^{0} \frac{1}{\mathbf{u}} \mathrm{du}+\int_{0}^{1} \frac{1}{\mathbf{u}} \mathrm{du}+\int_{1}^{-\ln x} \frac{1}{\mathfrak{u}} d \mathfrak{u}
$$

split, from which we get that expression (9) is equal to

$$
\gamma+\ln \ln \frac{1}{x}+\int_{0}^{-\ln x} \frac{e^{-u}}{u} d u+\int_{\ln x}^{0} \frac{1}{u} d u
$$

where interchanging the limits of the integration in the third term, and by applying the $u=-t$ substitution in the last term, one can get

$$
\gamma+\ln \ln \frac{1}{x}+\int_{-\ln x}^{0} \frac{1-e^{-u}}{u} d u
$$

where the integral can be exchanged with the sum given in the lemma, as in article [6]. If $x \in\left(e^{-1}, 1\right)$, then $0<-\ln x<1$, so one can do the

$$
\int_{\ln x}^{-\ln x} \frac{1}{u} d u=\int_{\ln x}^{0} \frac{1}{u} d u+\int_{0}^{1} \frac{1}{u} d u-\int_{-\ln x}^{1} \frac{1}{u} d u
$$

split, which we can transform back to the previous case by interchanging the limits of the integration in the last term.

Take note that when $x \in(0,1)$, then $\operatorname{li}(x)$ is smaller than zero and monotone decreases toward negative infinity, furthermore when $x \in(1,+\infty)$ then it monotone increases from negative infinity to positive infinity, see figure 2 .

The reason why we have given the results in proposition 1 by using the logarithmic integral, and not dissecting it further is because it is hard to give a concise and also precise approximation for the logarithmic integral with elementary functions.

$x$
Figure 2: Plot of the logarithmic integral. The curves between 0 and 1 correspond to the truncated versions of the equation from lemma 1. (They get darker as we take more terms from the sum.)

## 2 Proof of the proposition

As it is noted, from now on we are going to concentrate on the positive case of expression $(1)$, while $s \in[1 / 2,1)$ holds.
Proof. Changing the product into summation in expression (1), we get

$$
\begin{equation*}
\exp \left(\ln \prod_{\mathfrak{p} \leq n}\left(1+\frac{1}{\mathfrak{p}^{s}}\right)\right)=\exp \left(\sum_{p \leq n} \ln \left(1+\frac{1}{\mathfrak{p}^{s}}\right)\right) \tag{10}
\end{equation*}
$$

where we are going to apply theorem 1 on the argument of the exponential function on the right hand side. According theorem 1 and our observations after it, this sum is equal to

$$
\begin{equation*}
\int_{2}^{n} \frac{\ln \left(1+\frac{1}{t^{s}}\right)}{\ln t} d t+(\varepsilon(n)+\operatorname{li}(2)) \ln \left(1+\frac{1}{n^{s}}\right)+s \int_{2}^{n} \frac{\varepsilon(t)+\operatorname{li}(2)}{t\left(t^{s}+1\right)} d t \tag{11}
\end{equation*}
$$

because there exists an open interval containing $[2, \infty)$ on which we can continuously differentiate $\ln \left(1+1 / t^{s}\right)$, furthermore

$$
\frac{\mathrm{d}}{\mathrm{dt}} \ln \left(1+\frac{1}{\mathrm{t}^{s}}\right)=-\frac{\mathrm{s}}{\mathrm{t}\left(\mathrm{t}^{s}+1\right)}
$$

holds in the said interval. In the following sections we will examine the terms of expression (11) piecewise, then sum our results in section 2.4.

### 2.1 First term

Concerning the first term in expression (11), we are going to use a series representation of the logarithmic function from [1], equation 4.1.24, which goes as

$$
\begin{equation*}
\ln (1+x)=x-\frac{1}{2} x^{2}+\frac{1}{3} x^{3}+\ldots=\sum_{k=1}^{\infty} \frac{(-1)^{k+1} x^{k}}{k} \tag{12}
\end{equation*}
$$

where $|x| \leq 1$ and $x \neq-1$. So

$$
\begin{equation*}
\int_{2}^{n} \frac{\ln \left(1+\frac{1}{t^{s}}\right)}{\ln t} d t=\int_{2}^{n} \sum_{k=1}^{\infty} \frac{(-1)^{k+1}}{k t^{k s} \ln t} d t=\int_{2}^{n} \sum_{k=1}^{\infty} f_{k, s}(t) d t \tag{13}
\end{equation*}
$$

because $\left|1 / t^{s}\right| \leq 1$ when $t \in[2, n]$ and $s \in[1 / 2,1]$. Now we are going to show that the order of the integration and the summation can be interchanged in equation (13). For every $k>0$ and $s \in[1 / 2,1]$, we have that $f_{k, s}$ is continuous on $[2, n]$, which means that it is measurable on $[2, n]$. What we have to show is that

$$
\sum_{k=1}^{\infty} \int_{2}^{n}\left|f_{k, s}(t)\right| d t=\sum_{k=1}^{\infty} \int_{2}^{n} \frac{1}{k t^{k s} \ln t} d t<\frac{1}{\ln 2} \sum_{k=1}^{\infty} \int_{2}^{n} \frac{1}{t^{k s}} d t
$$

converges. Because the integrand on the right hand side is a positive, measurable function on $[2, n]$ for every $k>0$ and $s \in[1 / 2,1]$, we can interchange the order of the summation and the integration, which - based on the sum of the geometric series - gives us

$$
\frac{1}{\ln 2} \int_{2}^{n} \sum_{k=1}^{\infty} \frac{1}{t^{k s}} d t=\frac{1}{\ln 2} \int_{2}^{n} \frac{1}{t^{s}-1} d t \leq \frac{1}{\ln 2} \int_{2}^{n} \frac{1}{\sqrt{t}-1} d t<+\infty
$$

because $s \in[1 / 2,1]$, so one can interchange the order of summation and integration in equation (13) as

$$
\begin{equation*}
\int_{2}^{n} \sum_{k=1}^{\infty} \frac{(-1)^{k+1}}{k t^{k s} \ln t} d t=\sum_{k=1}^{\infty} \frac{(-1)^{k+1}}{k} \int_{2}^{n} \frac{1}{t^{k s} \ln t} d t \tag{14}
\end{equation*}
$$

where we have two cases during the evaluation of the integral inside the summation.

1. If $s=1 / \mathrm{m}$ for some $\mathrm{m}>0$ integer, then the integral on the right side in equation (14) will be simply $\ln \ln n-\ln \ln 2$ when $k=m$.
2. Otherwise, when $s \in(1 / 2,1)$ and there doesn't exist an $m$ integer such that $s=1 / m$, then the integral can be treated as follows. Let us set $t=r^{\lambda}$, then we get that

$$
\int_{2}^{n} \frac{1}{t^{k s} \ln t} d t=\int_{2^{1 / \lambda}}^{n^{1 / \lambda}} \frac{r^{\lambda-k s \lambda-1}}{\ln r} d r
$$

where we want $\lambda-k s \lambda-1$ to be zero, so we have to set $\lambda=1 /(1-k s)$, which gives us that

$$
\int_{2^{1 / \lambda}}^{\mathrm{n}^{1 / \lambda}} \frac{\mathrm{r}^{\lambda-k s \lambda-1}}{\ln r} \mathrm{dr}=\int_{2^{1-k s}}^{\mathrm{n}^{1-k s}} \frac{1}{\ln r} \mathrm{dr}=\operatorname{li}\left(\mathrm{n}^{1-\mathrm{ks}}\right)-\operatorname{li}\left(2^{1-\mathrm{ks}}\right)
$$

holds in this case.
Using these results, we get that when $s=1 / \mathrm{m}$ for some $\mathrm{m}>0$ integer, then equation (14) is equal to

$$
\begin{equation*}
\frac{(-1)^{m+1}}{m}(\ln \ln n-\ln \ln 2)+\sum_{\mathrm{k} \in \mathbb{N}^{+} \backslash\{\mathrm{m}\}} \frac{(-1)^{\mathrm{k}+1}}{\mathrm{k}}\left(\operatorname{li}\left(n^{1-\mathrm{ks}}\right)-\operatorname{li}\left(2^{1-\mathrm{ks}}\right)\right) \tag{15}
\end{equation*}
$$

otherwise when $s \in(1 / 2,1)$ and there doesn't exist an $m$ integer such that $s=1 / m$, then equation (14) is equal to the

$$
\begin{equation*}
\sum_{k=1}^{\infty} \frac{(-1)^{k+1}}{k}\left(\operatorname{li}\left(n^{1-k s}\right)-\operatorname{li}\left(2^{1-k s}\right)\right) \tag{16}
\end{equation*}
$$

sum. We are going to investigate these sums depending on the value of $s$ separately in the cases when $s \in(1 / 2,1)$, and when $s=1 / 2$.

### 2.1.1 Above half

When $s \in(1 / 2,1)$, then there is surely no such $m$ integer that $s=1 / m$, so we are going to concentrate on expression (16) in this case. We will show that the sum can be actually split into two sums; one which contains only the li( $\mathrm{n}^{1-\mathrm{ks}}$ ) terms, and another, which contains only the $\operatorname{li}\left(2^{1-k s}\right)$ terms. Considering the members of the first sum, when $1-\mathrm{ks}<0$, then

$$
\lim _{n \rightarrow \infty} \operatorname{li}\left(n^{1-k s}\right)=0
$$

holds. Because $s \in(1 / 2,1)$, this is true when $k>1$. Regarding the members of the second sum, because the logarithmic integral is negative on $(0,1)$, we have that

$$
\begin{equation*}
0<\frac{-\operatorname{li}\left(2^{1-\mathrm{ks}}\right)}{k}=-\frac{1}{k} \lim _{h \rightarrow 0^{+}} \int_{h}^{2^{1-k s}} \frac{1}{\ln t} d t \tag{17}
\end{equation*}
$$

holds for $k>2$ and $s \in(1 / 2,1)$. With these assumptions about $k$ and $s$, we have that the $1 / \ln t$ function is continuous on $\left[h, 2^{1-k s}\right]$ for every $h \in\left(0,2^{1-k s}\right)$, so we can apply the mean value theorem and get that the right hand side of equation (17) is smaller than

$$
\begin{equation*}
-\frac{1}{k} \lim _{\mathrm{h} \rightarrow 0^{+}} \frac{2^{1-\mathrm{ks}}-\mathrm{h}}{\ln 2^{1-\mathrm{ks}}}=\frac{2^{1-\mathrm{ks}}}{\mathrm{k}(\mathrm{ks}-1) \ln 2} \tag{18}
\end{equation*}
$$

because $|\ln t|$ increases as $t$ approaches zero from the right. Take note that the right hand side of equation (18) decreases monotonically to zero as $k$ approaches infinity when $s \in(1 / 2,1)$. To simplify the discussion, we introduce the

$$
\alpha_{i, s}(x):=\sum_{k=i}^{\infty} \frac{(-1)^{k+1}}{k} \operatorname{li}\left(x^{1-k s}\right)
$$

notation. We have arrived at the conclusion that $\alpha_{2, s}(n)$ converges to zero as $n$ approaches positive infinity, and $\alpha_{3, s}(2)$ also converges based on the alternating series test, so the sum in expression (16) can be split, and it is equal to

$$
\operatorname{li}\left(n^{1-s}\right)+\alpha_{2, s}(n)-\operatorname{li}\left(2^{1-s}\right)+\frac{1}{2} \operatorname{li}\left(2^{1-2 s}\right)+\mathcal{O}(1)
$$

when $s \in(1 / 2,1)$. Now we are going to bound the third and fourth terms in this last expression. Using equation (8) for the third term, we have

$$
\operatorname{li}\left(2^{1-s}\right)=\gamma+\ln \ln 2^{1-s}+\mathcal{O}(1)
$$

furthermore using the result of lemma 1 for the fourth term, we get

$$
\operatorname{li}\left(2^{1-2 s}\right)=\gamma+\ln \ln \frac{1}{2^{1-2 s}}+\mathcal{O}(1)
$$

where the

$$
\sum_{k=1}^{\infty} \frac{\ln ^{k} x}{k!k}<\sum_{k=0}^{\infty} \frac{\ln ^{k} x}{k!}=x
$$

inequality was applied in both cases. Summing these results, we have that expression (16) is equal to

$$
\begin{equation*}
\operatorname{li}\left(n^{1-s}\right)+\alpha_{2, s}(n)+\ln \frac{\sqrt{2 s-1}}{(1-s) \sqrt{\ln 2}}+\mathcal{O}(1) \tag{19}
\end{equation*}
$$

in this case. Now we keep the $\alpha_{2, s}(n)$ term, because it will disappear when we will let $n$ approach positive infinity at the end.

### 2.1.2 At half

As we lower $s$ from one toward zero, the $s=1 / 2$ is the first case where we have to use expression (15), which gives us

$$
\operatorname{li}(\sqrt{n})-\operatorname{li}(\sqrt{2})-\frac{1}{2}(\ln \ln n-\ln \ln 2)+\sum_{\mathrm{k}=3}^{\infty} \frac{(-1)^{\mathrm{k}+1}}{\mathrm{k}}\left(\operatorname{li}\left(\mathrm{n}^{1-\mathrm{k} / 2}\right)-\operatorname{li}\left(2^{1-\mathrm{k} / 2}\right)\right)
$$

where we can split the sum again, based on the arguments in section 2.1.1. Summing the constants, we get that expression (15) is equal to

$$
\begin{equation*}
\operatorname{li}(\sqrt{n})+\alpha_{3,1 / 2}(n)-\frac{1}{2} \ln \ln n+\mathcal{O}(1) \tag{20}
\end{equation*}
$$

when $s=1 / 2$. Yet again, the $\alpha_{3,1 / 2}(n)$ term will disappear when we will approach positive infinity with n at the end.

### 2.2 Second term

As for the second term in expression (11), we are going to use the inequalities 4.1.33 from [1], which state that

$$
\frac{x}{1+x}<\ln (1+x)<x
$$

holds for every $x>-1, x \neq 0$, from which it follows that

$$
\frac{1}{x^{s}+1}<\ln \left(1+\frac{1}{x^{s}}\right)<\frac{1}{x^{s}}
$$

holds for every $x \in(0, \infty)$ and $s \in \mathbb{R}$. Using the error term from inequality (4) in the second term of expression (11), we get that

$$
\left|(\varepsilon(n)+\operatorname{li}(2)) \ln \left(1+\frac{1}{n^{s}}\right)\right| \leq(|\varepsilon(n)|+\operatorname{li}(2)) \ln \left(1+\frac{1}{n^{s}}\right)
$$

$$
\begin{equation*}
<\frac{\sqrt{n}+\operatorname{li}(2)}{n^{s}} \tag{21}
\end{equation*}
$$

so if $s>1 / 2$, then the absolute value of the second term converges to zero as $n$ goes to infinity. When $s=1 / 2$, then its absolute value is smaller than

$$
\begin{equation*}
1+\frac{\operatorname{li}(2)}{\sqrt{n}} \tag{22}
\end{equation*}
$$

otherwise, its absolute value behaves asymptotically as

$$
\begin{equation*}
\mathcal{O}\left(n^{1 / 2-s}\right) \tag{23}
\end{equation*}
$$

when $s<1 / 2$.

### 2.3 Third term

For the third term in expression (11), we can assume without loss of generality that $\varepsilon$ and also $|\varepsilon|$ is Riemann-integrable on $[2, n]$, so we can write

$$
\begin{equation*}
\left|s \int_{2}^{n} \frac{\varepsilon(t)+\operatorname{li}(2)}{t\left(t^{s}+1\right)} d t\right| \leq s \int_{2}^{n} \frac{|\varepsilon(t)|+\operatorname{li}(2)}{t\left(t^{s}+1\right)} d t<s \int_{2}^{n} \frac{\sqrt{t}+\operatorname{li}(2)}{t\left(t^{s}+1\right)} d t \tag{24}
\end{equation*}
$$

where we have substituted the error term from inequality (4). This is equal to

$$
\begin{equation*}
\beta_{s}(n):=2 s\left[\sqrt{t} \cdot{ }_{2} F_{1}\left(1, \frac{1}{2 s} ; 1+\frac{1}{2 s} ;-t^{s}\right)\right]_{2}^{n}+\operatorname{li}(2)\left[\ln \frac{t^{s}}{t^{s}+1}\right]_{2}^{n} \tag{25}
\end{equation*}
$$

because the first part of the integral on the right hand side can be transformed into the form of the ${ }_{2} \mathrm{~F}_{1}$ Gauss hypergeometric function, and the second part can be decomposed into partial fractions. The transformation of the first part can be done as

$$
\begin{aligned}
\int \frac{1}{\sqrt{x}\left(x^{s}+1\right)} d x & =\int_{0}^{x} u^{-\frac{1}{2}}\left(1+u^{s}\right)^{-1} d u+C \\
& =\frac{1}{s} \int_{0}^{x^{s}} r^{\frac{1}{2 s}-1}(1+r)^{-1} d r+C \\
& =\frac{\sqrt{x}}{s} \int_{0}^{1} t^{\frac{1}{2 s}-1}\left(1+x^{s} t\right)^{-1} d t+C
\end{aligned}
$$

where first we switched the indefinite integral into a definite one, then we have applied the $u^{s}=r$ substitution and finally the $r=\chi^{s} t$ substitution. The Gauss hypergeometric function can be written in the

$$
{ }_{2} F_{1}(a, b ; c ; z)=\frac{\Gamma(c)}{\Gamma(b) \Gamma(c-b)} \int_{0}^{1} t^{b-1}(1-t)^{c-b-1}(1-t z)^{-a} d t
$$

form, see equation 15.3.1 in [1]. Using this, we have that $z=-\chi^{s}, a=1$ and $\mathrm{b}=1 / 2 \mathrm{~s}$. Because $\mathrm{c}-\mathrm{b}-1=0$ should hold, we get our missing c and with it, the first part of expression (25). As for the second part, we have

$$
\operatorname{li}(2) \int_{2}^{n} \frac{s}{\mathfrak{t}\left(\mathrm{t}^{s}+1\right)} \mathrm{dt}=\operatorname{li}(2) \int_{2}^{n} \frac{s}{\mathrm{t}}-\frac{s \mathrm{t}^{s-1}}{\mathrm{t}^{s}+1} d \mathrm{t}
$$

where the second fraction is a logarithmic derivative. We are going to investigate the resulting expression (25) separately in the cases when $s \in(1 / 2,1)$, and when $s=1 / 2$.

### 2.3.1 Above half

First, we deal with the upper limit of the integration. Substituting $n$ into expression (25), we get

$$
\begin{equation*}
2 s \sqrt{n} \cdot{ }_{2} F_{1}\left(1, \frac{1}{2 s} ; 1+\frac{1}{2 s} ;-n^{s}\right)+\operatorname{li}(2) \ln \frac{n^{s}}{n^{s}+1} \tag{26}
\end{equation*}
$$

which, by using the

$$
{ }_{2} \mathrm{~F}_{1}(\mathrm{a}, \mathrm{~b} ; \mathrm{c} ; z)=(1-z)^{-\mathrm{b}}{ }_{2} \mathrm{~F}_{1}\left(\mathrm{~b}, \mathrm{c}-\mathrm{a} ; \mathrm{c} ; \frac{z}{z-1}\right)
$$

linear transformation formula, see equation 15.3 .5 in [1], can be transformed into the

$$
2 s \frac{\sqrt{n}}{\left(n^{s}+1\right)^{1 / 2 s}} 2_{1}\left(\frac{1}{2 s}, \frac{1}{2 s} ; 1+\frac{1}{2 s} ; \frac{n^{s}}{n^{s}+1}\right)+\operatorname{li}(2) \ln \frac{n^{s}}{n^{s}+1}
$$

form. Because $s \in(1 / 2,1)$, we get that

$$
\lim _{n \rightarrow+\infty} \frac{\sqrt{n}}{\left(n^{s}+1\right)^{1 / 2 s}} 2 F_{1}\left(\frac{1}{2 s}, \frac{1}{2 s} ; 1+\frac{1}{2 s} ; \frac{n^{s}}{n^{s}+1}\right)={ }_{2} F_{1}\left(\frac{1}{2 s}, \frac{1}{2 s} ; 1+\frac{1}{2 s} ; 1\right)
$$

holds. When $c$ is not zero or a negative integer, furthermore $\mathfrak{R}(c-a-b)>0$ is true, then one can apply the

$$
{ }_{2} \mathrm{~F}_{1}(\mathrm{a}, \mathrm{~b} ; \mathrm{c} ; 1)=\frac{\Gamma(\mathrm{c}) \Gamma(\mathrm{c}-\mathrm{a}-\mathrm{b})}{\Gamma(\mathrm{c}-\mathrm{a}) \Gamma(\mathrm{c}-\mathrm{b})}
$$

substitution, see equation 15.1 .20 in [1]. Because $s \in(1 / 2,1)$, the conditions are satisfied, and we can utilise the above mentioned formula to get

$$
\begin{equation*}
2 s_{2} F_{1}\left(\frac{1}{2 s}, \frac{1}{2 s} ; 1+\frac{1}{2 s} ; 1\right)=2 s \Gamma\left(1+\frac{1}{2 s}\right) \Gamma\left(1-\frac{1}{2 s}\right) \tag{27}
\end{equation*}
$$

which has an anomaly at $s=1 / 2$. For the second term, observe that

$$
\lim _{n \rightarrow+\infty} \ln \frac{n^{s}}{n^{s}+1}=0
$$

holds. Now we deal with the lower limit of the integral. By substituting 2 into expression (25) we get

$$
\begin{equation*}
2 \mathrm{~s} \sqrt{2} \cdot{ }_{2} \mathrm{~F}_{1}\left(1, \frac{1}{2 \mathrm{~s}} ; 1+\frac{1}{2 \mathrm{~s}} ;-2^{s}\right)+\operatorname{li}(2) \ln \frac{2^{s}}{2^{s}+1} \tag{28}
\end{equation*}
$$

where the second term is a small constant, so we will concentrate on the value of the hypergeometric function. Using another

$$
{ }_{2} \mathrm{~F}_{1}(\mathrm{a}, \mathrm{~b} ; \mathrm{c} ; z)=(1-z)^{-a}{ }_{2} \mathrm{~F}_{1}\left(\mathrm{a}, \mathrm{c}-\mathrm{b} ; \mathrm{c} ; \frac{z}{z-1}\right)
$$

linear transformation formula, see equation 15.3.4 in [1], we get that the hypergeometric function in expression (28) is equal to

$$
\frac{1}{2^{s}+1^{2}}{ }^{2} F_{1}\left(1,1 ; 1+\frac{1}{2 s} ; \frac{2^{s}}{2^{s}+1}\right)
$$

where the last argument is in $(0,1)$. We are going to give an upper bound for this expression. If $a \leq 1$ and $0<b \leq c$, then

$$
{ }_{2} \mathrm{~F}_{1}(-\mathrm{a}, \mathrm{~b} ; \mathrm{c} ; z)^{1 / a} \geq\left[\left(1-\frac{\mathrm{b}}{\mathrm{c}}\right)+\frac{\mathrm{b}}{\mathrm{c}}(1-z)^{\mathrm{a}}\right]^{1 / a}
$$

for all $z \in(0,1)$, see $[8]$ furthermore [13] and [4]. Because $s \in(1 / 2,1)$, we can apply this inequality, which means that

$$
{ }_{2} \mathrm{~F}_{1}\left(1,1 ; 1+\frac{1}{2 \mathrm{~s}} ; \frac{2^{s}}{2^{s}+1}\right) \leq\left(1-\frac{1}{1+\frac{1}{2 s}}\right)+\frac{1}{1+\frac{1}{2 s}}\left(1-\frac{2^{s}}{2^{s}+1}\right)^{-1}
$$

which is just a small constant when $s \in(1 / 2,1)$. Joining our results so far, we get that $\beta_{s}(\mathfrak{n})$ from expression (25) then converges to the right hand side of equation (27) plus some constant as $n$ approaches positive infinity when $s \in(1 / 2,1)$.

### 2.3.2 At half

One of the special elementary cases of the hypergeometric function is the

$$
{ }_{2} F_{1}(1,1 ; 2 ; z)=-z^{-1} \ln (1-z)
$$

equality, see equation 15.1.3 in [1]. This means that when $s=1 / 2$, then the first part of expression (25) is equal to

$$
\begin{equation*}
[\ln (\sqrt{t}+1)]_{2}^{n} \in \mathcal{O}(\ln n) \tag{29}
\end{equation*}
$$

and because we have tried to bound the absolute value of the third term from above, this partial result already spoils our pursuit of reaching a constant error term in this special case. Regarding the second part of expression (25), what was said in the previous section still holds, meaning that the second part converges to a small constant as $n$ approaches infinty.

### 2.4 Summing the parts

Now we are going to sum our results. In the case when $s \in(1 / 2,1)$, the first term of expression (11) is equal to expression (19), the absolute value of the second term is smaller than the right hand side of inequality (21) and the absolute value of the third term is smaller than expression (25). So expression (11) is equal to

$$
\operatorname{li}\left(n^{1-s}\right)+\ln \frac{\sqrt{2 s-1}}{(1-s) \sqrt{\ln 2}}+\mathcal{O}(1)+\mathcal{O}\left(\alpha_{2, s}(n)+\frac{\sqrt{n}+\operatorname{li}(2)}{n^{s}}+\beta_{s}(n)\right)
$$

in this case. Reintroducing this into equation (10), we get that the positive case of expression (1) is equal to

$$
\frac{\sqrt{2 s-1}}{(1-s) \sqrt{\ln 2}} \exp \left(\operatorname{li}\left(n^{1-s}\right)+\mathcal{O}(1)+\mathcal{O}\left(\alpha_{2, s}(n)+\frac{\sqrt{n}+\operatorname{li}(2)}{n^{s}}+\beta_{s}(n)\right)\right)
$$

where, after dividing with $\exp \left(\operatorname{li}\left(n^{1-s}\right)\right)$ and taking the limit in $n$, we get the sought equality (5). As for the case when $s=1 / 2$ the first term of expression (11) is equal to expression (20), the absolute value of the second term is smaller than expression (22), and the absolute value of the third term is smaller than expression (29). By these, expression (11) is equal to

$$
\operatorname{li}(\sqrt{n})+\mathcal{O}(\ln n)
$$

and by reintroducing this into equation (10), we get equality (7).

## 3 Remarks

### 3.1 Below half

The problem in this region is that either when using expression (15) or expression (16) the result becomes more and more complicated as we approach zero with $s$ from the right. This happens mainly because the li $\left(n^{1-k s}\right)$ terms only disappear when $\mathrm{k}>1 / \mathrm{s}$, but also because one has to pay close attention to the $\mathrm{li}\left(2^{1-\mathrm{ks}}\right)$ terms when $2^{1-\mathrm{ks}}$ is close to one. (The $\mathrm{li}\left(\mathrm{n}^{1-\mathrm{ks}}\right)$ terms behave more nicely, because they avoid the anomaly at one.) Furthermore, without taking the third term into consideration, the second term already contributes a rough asymptotical term, see expression (23).

### 3.2 The case of negative exponents

When the $s$ exponent is negative, instead of expression (11), the sum in equation (10) is equal to

$$
\begin{equation*}
\int_{2}^{\mathrm{n}} \frac{\ln \left(1+\mathrm{t}^{|s|}\right)}{\ln \mathrm{t}} \mathrm{dt}+(\varepsilon(\mathrm{n})+\operatorname{li}(2)) \ln \left(1+\mathrm{n}^{|s|}\right)-|s| \int_{2}^{\mathrm{n}} \frac{\varepsilon(\mathrm{t})+\operatorname{li}(2)}{\mathrm{t}^{1-|s|}\left(1+\mathrm{t}^{|s|}\right)} \mathrm{dt} \tag{30}
\end{equation*}
$$

because there exists an open interval containing $[2, \infty)$ on which we can continuously differentiate $\ln \left(1+t^{|s|}\right)$, furthermore

$$
\frac{\mathrm{d}}{\mathrm{dt}} \ln \left(1+\mathrm{t}^{|s|}\right)=|s| \frac{\mathrm{t}^{|s|-1}}{1+\mathrm{t}^{|s|}}
$$

holds. We cannot apply equation (12) like in section 2.1. Instead, we will use the following simple estimations, which can be shown using the properties of the logarithmic function. For every $s>0$ real number, there exists such $c_{s}>1$ constant, that

$$
\ln \chi^{s} \leq \ln \left(1+\chi^{s}\right) \leq c_{s} \ln \chi^{s}
$$

holds for every $x>1$ real number. Applying these estimations on the first term on expression (30), we get

$$
|s| \int_{2}^{n} d t \leq \int_{2}^{n} \frac{\ln \left(1+t^{|s|}\right)}{\ln t} d t \leq c_{s}|s| \int_{2}^{n} d t
$$

which shows us that the first term grows as $\Theta(n)$. Proceeding to the second term in expression (30), we have that its absolute value is smaller than

$$
(|\varepsilon(n)|+\operatorname{li}(2)) \ln \left(1+n^{|s|}\right) \leq c_{s}|s|(\sqrt{n}+\operatorname{li}(2)) \ln n
$$

which grows as $\mathcal{O}(\sqrt{n} \ln n)$. Still assuming that $\varepsilon$ and also $|\varepsilon|$ is Riemannintegrable on $[2, n]$, the absolute value of the third term in expression (30) is smaller than or equal to

$$
|s| \int_{2}^{n} \frac{|\varepsilon(\mathrm{t})|+\mathrm{l}(2)}{\mathrm{t}^{1-|s|}\left(1+\mathrm{t}^{|s|}\right)} \mathrm{dt}
$$

which we can estimate from above by dropping the plus one from the denominator. This way, we get

$$
|s| \int_{2}^{n} \frac{|\varepsilon(t)|+\operatorname{li}(2)}{t} d t<|s| \int_{2}^{n} \frac{\sqrt{\mathrm{t}}+\operatorname{li}(2)}{\mathrm{t}} \mathrm{dt}=|\mathrm{s}|[2 \sqrt{\mathrm{t}}+\operatorname{li}(2) \ln \mathrm{t}]_{2}^{\mathrm{n}}
$$

which grows as $\mathcal{O}(\sqrt{n})$. As it can be seen, when the $s$ exponent is negative in equation (10), then the first term dominates so the positive case of the product in expression (1) grows as $\exp (\Theta(n))$.

### 3.3 Remark about the strength of the method

Finally, we are going to look at how the theorem performs in the $s=1$ case. Following the same path as in section 2, for the first term in expression (11) we should use expression (15), which is equal to

$$
\ln \ln n-\ln \ln 2+\sum_{k=2}^{\infty} \frac{(-1)^{k+1}}{k}\left(\operatorname{li}\left(n^{1-k}\right)-\operatorname{li}\left(2^{1-k}\right)\right)
$$

in this case. As in section 2.1.1, we can deduce that this expression is equal to

$$
\begin{equation*}
\ln \ln n+\alpha_{2,1}(n)+\mathcal{O}(1) \tag{31}
\end{equation*}
$$

where the $\alpha_{2,1}(n)$ term disappears as we approach positive infinity with $n$. The absolute value of the second term in expression (11) is smaller than

$$
\begin{equation*}
\frac{\sqrt{n}+\operatorname{li}(2)}{n} \tag{32}
\end{equation*}
$$

which vanishes as $n$ tends to positive infinity, so what remains is the absolute value of the third term in the expression (11). Based on inequality (24) we get that its absolute value is smaller than

$$
\beta(n):=\int_{2}^{n} \frac{\sqrt{\mathrm{t}}+\operatorname{li}(2)}{\mathrm{t}(\mathrm{t}+1)} \mathrm{dt}=2[\arctan \sqrt{\mathrm{t}}]_{2}^{\mathrm{n}}+\left[\ln \frac{\mathrm{t}}{\mathrm{t}+1}\right]_{2}^{\mathrm{n}}
$$

which converges to a small constant when $\mathfrak{n}$ approaches infinity. Using expression (31), expression (32), and finally expression (33), we get that equation (10) is equal to

$$
\exp \left(\ln \ln n+\mathcal{O}(1)+\mathcal{O}\left(\alpha_{2,1}(n)+\frac{\sqrt{n}+\operatorname{li}(2)}{n}+\beta(n)\right)\right)
$$

which in turn means that

$$
\lim _{n \rightarrow \infty} \frac{1}{\ln n} \prod_{p \leq n}\left(1+\frac{1}{p}\right)=e^{\mathcal{O}(1)}
$$

holds.

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## References

[1] M. Abramowitz, I. A. Stegun, Handbook of Mathematical Functions with Formulas, Graphs, and Mathematical Tables, Applied Mathematics Series 55 , tenth printing, Dover publications, (1972).
[2] T. M. Apostol Introduction to Analytic Number Theory, Undergraduate Texts in Mathematics, Springer-Verlag, (1976).
[3] E. Bach, J. Shallit, Algorithmic Number Theory, Volume 1: Efficient Algorithms, The MIT Press, (1996).
[4] R. W., Barnard, K. C. Richards, A Note on the Hypergeometric Mean Value, Comput. Meth. Funct. Th., 1 (1) (2001), 81-88.
[5] B. C. Berndt, Ramanujan's Notebooks, Part I., Springer, (1985).
[6] B. C. Berndt, R. J. Evans, Some elegant approximations and asymptotic formulas of Ramanujan, J. Comput. Appl. Math., 37 (1991), 35-41.
[7] R. P. Brent, On the Zeros of the Riemann Zeta Function in the Critical Strip. Math. Comp., 33 (148) (1979), 1361-1372.
[8] B. C. Carlson, Some inequalities for hypergeometric functions. Proc. Amer. Math. Soc., 17 (1966), 32-39.
[9] L. Euler, Variae observationes circa series infinitas. Commentarii academiae scientiarum Petropolitanae, 9 (1737), 160-188.
[10] H. Koch, Sur la distribution des nombres premiers. Acta Math., 24 (1901), 159-182.
[11] T. Kotnik, The prime-counting function and its analytic approximations, $\pi(x)$ and its approximations. Adv. Comput. Math., 29 (1) (2008), 55-70.
[12] F. Mertens, Ein Beitrag zur analytischen Zahlentheorie, J. reine angew. Math., 78 (1874), 46-62.
[13] K. C. Richards, Sharp power mean bounds for the Gaussian hypergeometric function. J. Math. Anal. Appl., 308 (1) (2005), 303-313.
[14] L. Schoenfeld, Sharper Bounds for the Chebyshev Functions $\theta(x)$ and $\psi(x)$. II. Math. Comp., 30 (134) (1976), 337-360.

# More on decomposition of generalized continuity 

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#### Abstract

In this paper a new class of sets termed as $\omega_{\mu}^{*}$-open sets has been introduced and studied. Using these concept, a unified theory for decomposition of $(\mu, \lambda)$-continuity has been given.


## 1 Introduction

For the last one decade or so, the notion of generalized topological spaces and several classes of generalized types of open sets are being studied by different mathematicians. Our aim here is to study the notion of decomposition of continuity by using the concept of generalized topology introduced by Á. Császár [2]. On the otherhand the notion of decompositions of continuity was first introduced by Tong $[18,19]$ by defining $\mathcal{A}$-sets and $\mathcal{B}$-sets. After then decompositions of continuity and some of its weak forms have been studied by Ganster and Reilly [7, 8], Yalvac [20], Hatir and Noiri [10, 11], Przemski [14], Noiri and Sayed [13], Dontchev and Przemski [5], Erguang and Pengfei [6] and many others. Decompositions of regular open sets and regular closed sets are given by using PS-regular sets in [9]. Since then the notion of decompositions of continuity is one of the most important area of research.

[^8]We first recall some definitions given in [2]. Let $X$ be a non-empty set and $\exp X$ denote the power set of $X$. We call a class $\mu \subseteq \exp X$ a generalized topology (briefly, GT) [1, 2], if $\varnothing \in \mu$ and $\mu$ is closed under arbitrary unions. A set $X$, with a GT $\mu$ on it is said to be a generalized topological space (briefly, GTS) and is denoted by $(X, \mu)$. A GT $\mu$ is said to be a quasi topology (briefly QT) [3] if $M, M^{\prime} \in \mu$ implies $M \cap M^{\prime} \in \mu$. The pair $(X, \mu)$ is said to be a QTS if $\mu$ is a QT on $X$. For a GTS $(X, \mu)$, the elements of $\mu$ are called $\mu$-open sets and the complements of $\mu$-open sets are called $\mu$-closed sets. A GTS $(X, \mu)$ is called a $\mu$-space [13] or a strong GTS [4] if $X \in \mu$. A subset $A$ of a topological space $(X, \tau)$ is called $\omega$-closed [12] if it contains all its condensation points. The complement of an $\omega$-closed set is called an $\omega$-open set. It is well known that a subset $A$ of a space $(X, \tau)$ is $\omega$-open if and only if for each $x \in A$, there exists $U \mathcal{U} \in \tau$ containing $x$ such that $U \backslash A$ is countable.

The purpose of this paper is to introduce the decomposition theorem for the $(\mu, \lambda)$ continuous functions introduced in [1] which is a generalization of continuity and different weak forms of continuity. Throughout the paper, by $(X, \mu)$ and $(Y, \lambda)$ we shall mean GTS unless otherwise stated.

## $2 \quad \omega_{\mu}^{*}$-open sets and its properties

Definition 1 Let $(\mathrm{X}, \mu)$ be a GTS. A subset A of X is called an $\omega_{\mu}^{*}$-open ( $\omega_{\mu}$-open [15]) set if for each $\chi \in A$, there exists a $\mu$-open set $U$ containing $\chi$ such that $\mathrm{U} \backslash \mathfrak{i}_{\mu}(\mathcal{A})$ (resp. $\mathrm{U} \backslash \mathcal{A}$ ) is countable. The complement of an $\omega_{\mu}^{*}$-open (resp. $\omega_{\mu}$-open) set is known as an $\omega_{\mu}^{*}$-closed (resp. $\omega_{\mu}$-closed [15]) set.

It follows from Definition 1 that every $\omega_{\mu}^{*}$-open set is an $\omega_{\mu}$-open set and every $\mu$-open set is $\omega_{\mu}^{*}$-open set but the converses are false as shown in Example 3.

Remark 1 Let $\mu$ be a GT on a topological space (X, $\tau)$. If $\tau \subseteq \mu$, then the following relations hold:

$$
\begin{array}{cc}
\begin{array}{c}
\omega \text {-open } \text { set }
\end{array} \Leftarrow \text { open set } \Rightarrow & \mu \text {-open set } \\
\Downarrow \\
\Downarrow \\
\omega_{\mu} \text {-open set } & \Leftarrow \\
\omega_{\mu}^{*} \text {-open set }
\end{array}
$$

Example 1 (a) Let $X=\mathbb{R}$, $\tau$ be the usual topology on $\mathbb{R}$. Let $\mu=\{\varnothing, X, \mathbb{Q}\}$. Then $\mu$ is a $G T$ on the topological space $(\mathrm{X}, \tau)$. It is easy to see that $\mathbb{Q}$ is an $\omega_{\mu}^{*}$-open set but not an $\omega$-open set.
(b) Let $\mathrm{X}=\mathbb{R}$ and $\mu$ be the usual topology on $\mathbb{R}$. Then $\mu$ is a GT on X . It is easy to see that $\mathbb{I}$, the set of irrationals is an $\omega_{\mu}$-open set but not an $\omega_{\mu}^{*}$-open set.
(c) Let $X=\mathbb{R}$ and $\mu=\{A \subseteq X: 0 \in A\} \cup\{\varnothing\}$. Then $\mu$ is a $G T$ on the set $X$. It is easy to see that $\mathbb{I}$, the set of irrationals is $\omega_{\mu}$-open but not $\mu$-open.
(d) Let $X=\mathbb{R}, \mu=\{A: A$ is uncountable $\} \cup\{\varnothing\}$ and $\tau=\{\varnothing, X, \mathbb{Q}\}$. Then $\mu$ is a $G T$ on the topological space $(\mathrm{X}, \tau)$. It can be easily verified that $\mathbb{Q}$ is an $\omega$-open set but not an $\omega_{\mu}^{*}$-open set.

The family of all $\omega_{\mu}^{*}$-open sets of a $\operatorname{GTS}(X, \mu)$ is denoted by $\omega_{\mu}^{*}(X)$ or simply by $\omega_{\mu}^{*}$.

Proposition 1 (a) In a GTS $(\mathrm{X}, \mu)$, the collection of all $\omega_{\mu}^{*}$-open sets forms a GT on X .
(b) If $(\mathrm{X}, \mu)$ is a $Q T S$, then the collection of all $\omega_{\mu}^{*}$-open sets forms a $Q T$ on X.

Proof. (a) It is obvious that $\varnothing$ is an $\omega_{\mu}^{*}$-open set. Let $\left\{A_{\alpha}: \alpha \in \Lambda\right\}$ be a collection of $\omega_{\mu}^{*}$-open subsets of $X$. Then for each $x \in \cup\left\{A_{\alpha}: \alpha \in \Lambda\right\}$, $x \in A_{\alpha}$ for some $\alpha \in \Lambda$. Thus there exists $U \in \mu$ containing $x$ such that $\mathrm{U} \backslash \mathfrak{i}_{\mu}\left(\mathrm{A}_{\alpha}\right)$ is countable. Now as $\mathrm{U} \backslash \mathfrak{i}_{\mu}\left(\cup\left\{\mathcal{A}_{\alpha}: \alpha \in \Lambda\right\}\right) \subseteq \mathrm{U} \backslash \mathfrak{i}_{\mu}\left(\mathrm{A}_{\alpha}\right)$, thus $\mathrm{U} \backslash \mathfrak{i}_{\mu}\left(\cup\left\{\mathcal{A}_{\alpha}: \alpha \in \Lambda\right\}\right)$ is countable. Hence $\cup\left\{A_{\alpha}: \alpha \in \Lambda\right\}$ is an $\omega_{\mu}^{*}$-open set.
(b) It follows from (a) that ( $X, \omega_{\mu}^{*}$ ) is a GTS. Let A and B be two $\omega_{\mu}^{*}$-open sets and $x \in A \cap B$. Then there exist $\mu$-open sets $U$ and $V$ containing $x$ such that $\mathrm{U} \backslash \mathfrak{i}_{\mu}(A)$ and $V \backslash \mathfrak{i}_{\mu}(B)$ are countable. Then $\mathrm{U} \cap \mathrm{V}$ is a $\mu$-open set containing $x$ and $(U \cap V) \backslash i_{\mu}(A \cap B)=(U \cap V) \backslash i_{\mu}(A) \cap \mathfrak{i}_{\mu}(B) \cong\left[U \backslash i_{\mu}(A)\right] \cup\left[V \backslash i_{\mu}(B)\right]$. Thus $(U \cap V) \backslash i_{\mu}(A \cap B)$ is countable so that $\omega_{\mu}^{*}$ is a QT on $X$.

Theorem 1 A subset A of a GTS $(X, \mu)$ is an $\omega_{\mu}^{*}$-open set if and only if for each $\mathrm{x} \in A$, there exist $\mathrm{U}_{\mathrm{x}} \in \mu$ containing x and a countable subset C such that $\mathrm{u}_{\mathrm{x}} \backslash \mathrm{C} \cong \mathfrak{i}_{\mu}(A)$.

Proof. Let $\mathcal{A}$ be an $\omega_{\mu}^{*}$-open set in $X$ and $x \in A$. Then there exists $\mathrm{U}_{\mathrm{x}} \in \mu$ containing $x$ such that $u_{x} \backslash i_{\mu}(A)$ is countable. Let $C=U_{x} \backslash i_{\mu}(A)=U_{x} \cap(X \backslash$ $\left.\mathfrak{i}_{\mu}(A)\right)$. Then $\mathrm{U}_{x} \backslash C \subseteq \mathfrak{i}_{\mu}(A)$.

Conversely, let $x \in A$ and there exist $\mathrm{U}_{x} \in \mu$ containing $x$ and a countable subset $C$ such that $\mathrm{U}_{x} \backslash \mathrm{C} \subseteq \mathfrak{i}_{\mu}(A)$. Then $\mathrm{U}_{\star} \backslash \mathfrak{i}_{\mu}(A) \subseteq C$ and hence $\mathrm{U}_{\star} \backslash \mathfrak{i}_{\mu}(A)$ is a countable set. Thus $A$ is an $\omega_{\mu}^{*}$-open set in $X$.

Theorem 2 Let $(\mathrm{X}, \mu)$ be a $G T S$ and $\mathrm{C} \cong \mathrm{X}$. If C is $\omega_{\mu}^{*}$-closed, then $\mathrm{C} \subseteq \mathrm{K} \cup \mathrm{B}$ for some $\mu$-closed set K and a countable subset B .

Proof. If $C$ be $\omega_{\mu}^{*}$-closed, then $X \backslash C$ is $\omega_{\mu}^{*}$-open and hence for each $x \in X \backslash C$, there exist $\mathrm{U} \in \mu$ containing x and a countable subset B such that $\mathrm{U} \backslash \mathrm{B} \subseteq$ $\mathfrak{i}_{\mu}(X \backslash C)=X \backslash c_{\mu}(C)$. Thus $c_{\mu}(C) \subseteq X \backslash(U \backslash B)=X \backslash(U \cap(X \backslash B))=(X \backslash U) \cup B$. Let $K=X \backslash U$. Then $K$ is $\mu$-closed such that $C \subseteq K \cup B$.

Proposition 2 In a $G T S(\mathrm{X}, \mu)$, $\omega_{\mu}^{*}=\omega_{\omega_{\mu}^{*}}^{*}$, where $\omega_{\mu}^{*}$ denotes the family of $\omega_{\mu}^{*}$-open sets of the $G T S(X, \mu)$.

Proof. By Remark 1, we have $\omega_{\mu}^{*} \subseteq \omega_{\omega_{\mu}^{*}}^{*}$. Let $A \in \omega_{\omega_{\mu}^{*}}^{*}$. Then for each $x \in A$, there exist $U_{x} \in \omega_{\mu}^{*}$ with $x \in U_{x}$ and a countable set $C_{x}$ such that $\mathrm{U}_{x} \backslash \mathrm{C}_{x} \subseteq \mathfrak{i}_{\mu}(A)$. Furthermore there exist a $\mathrm{V}_{x} \in \mu$ with $x \in \mathrm{~V}_{x}$ and a countable set $D_{x}$ such that $V_{x} \backslash D_{x} \subseteq i_{\mu}\left(U_{x}\right)$. Thus $V_{x} \backslash\left(C_{x} \cup D_{x}\right)=\left(V_{x} \backslash D_{x}\right) \backslash C_{x} \subseteq$ $u_{x} \backslash C_{x} \subseteq i_{\mu}(A)$. Since $C_{x} \cup D_{x}$ is a countable set, we obtain $A \in \omega_{\mu}^{*}$.

Remark 2 If $(\mathrm{X}, \mu)$ be a $\mu$-space, then $\left(\mathrm{X}, \omega_{\mu}^{*}\right)$ is an $\omega_{\mu}^{*}$-space.
Definition $2 A$ subset $A$ of $a \operatorname{GTS}(X, \mu)$ is called an (i) $\left(\omega_{\mu}, \omega\right)$-set if $i_{\omega_{\mu}^{*}}(A)=i_{\omega_{\mu}}(A)$.
(ii) $\left(\omega_{\mu}, \mu\right)$-set if $i_{\omega_{\mu}^{*}}(A)=i_{\mu}(A)$.

Remark 3 Every $\omega_{\mu}^{*}$-open set is an $\left(\omega_{\mu}, \omega\right)$-set and every $\mu$-open set is an $\left(\omega_{\mu}, \mu\right)$-set but the converses are usually not true.

Example 2 (a) Let $\mathrm{X}=\mathbb{R}, \mu=\{\varnothing, \mathbb{I}, X\}$. Then $\mu$ is a $G T$ on $X$. It can be checked that $\mathbb{N}\left(=\right.$ the set of natural numbers) is not an $\omega_{\mu}^{*}$-open but an $\left(\omega_{\mu}, \omega\right)$-set.
(b) Let $\mathrm{X}=\mathbb{R}, \mu=\{\varnothing,(2,3), X\}$. Then $\mu$ is a $G T$ on $X$. The set $\left(1, \frac{3}{2}\right)$ is an $\left(\omega_{\mu}, \mu\right)$-set but not a $\mu$-open set.

Theorem 3 A subset $A$ of a $G T S(X, \mu)$ is $\omega_{\mu}^{*}$-open if and only if $A$ is $\omega_{\mu}$ open and an $\left(\omega_{\mu}, \omega\right)$-set.

Proof. Since every $\omega_{\mu}^{*}$-open set is $\omega_{\mu}$-open (from definition) and an $\left(\omega_{\mu}, \omega\right)$ set (by Remark 3), we have nothing to show.

Conversely, let $A$ be an $\omega_{\mu}$-open and an $\left(\omega_{\mu}, \omega\right)$-set. Then $A=\mathfrak{i}_{\omega_{\mu}}(A)=$ $i_{\omega_{\mu}^{*}}(A)$. Thus $A$ is an $\omega_{\mu}^{*}$-open set.

Theorem 4 A subset $\mathcal{A}$ of a $G T S(X, \mu)$ is $\mu$-open if and only if $A$ is $\omega_{\mu}^{*}$-open and an $\left(\omega_{\mu}, \mu\right)$-set.

Proof. One part follows from the fact that every $\mu$-open set is $\omega_{\mu}^{*}$-open and $\left(\omega_{\mu}, \mu\right)$-set.

Conversely, let $A$ be an $\omega_{\mu}^{*}$-open set and an $\left(\omega_{\mu}, \mu\right)$-set. Then $A=i_{\omega_{\mu}^{*}}(A)=$ $\mathfrak{i}_{\mu}(A)$. Thus $A$ is $\mu$-open.

Definition 3 A GTS $(\mathrm{X}, \mu)$ is called
(i) $\mu$-locally countable if each $x \in X$ is contained in a countable $\mu$-open set.
(ii) anti $\mu$-locally countable if each non-empty $\mu$-open subsets are uncountable.

Theorem 5 Let $(X, \mu)$ be a $\mu$-locally countable space. Then
(i) for any subset $A$ of $X, A$ is $\omega_{\mu}^{*}$-open.
(ii) $A$ is $\omega_{\mu}^{*}$-open if and only if $A^{\mu}$ is $\omega_{\mu}$-open.

Proof. (i) Let $A \subseteq X$ and $x \in A$. Then there is a countable $\mu$-open set $U$ containing $x$. Then $U \backslash \mathfrak{i}_{\mu}(A)$ is a countable set. Thus $A$ is $\omega_{\mu}^{*}$-open.
(ii) One part follows from the fact that every $\omega_{\mu}^{*}$-open is $\omega_{\mu}$-open.

Conversely, let $\mathcal{A}$ be $\omega_{\mu}$-open. Then by (i), $\mathcal{A}$ is $\omega_{\mu}^{*}$-open.
Remark 4 Let $(\mathrm{X}, \mu)$ be a countable GTS. Then for any subset A of $\mathrm{X}, \mathcal{A}$ is $\omega_{\mu}^{*}$-open.

Theorem 6 (i) Let $(X, \mu)$ be a GTS and $A \subseteq X$. If $(X, \mu)$ is anti $\mu$-locally countable, then so is $\left(\mathrm{X}, \omega_{\mu}^{*}\right)$.
(ii) Let $(X, \mu)$ be a QTS which is anti $\mu$-locally countable. Then for any $\mu$-open subset $A$ of $X, c_{\mu}(A)=c_{\omega_{\mu}^{*}}(A)$.

Proof. (i) Let $A$ be an $\omega_{\mu}^{*}$-open set and $x \in A$. Then there exist $U_{x} \in \mu$ containing $x$ and a countable subset $C$ such that $u_{x} \backslash C \subseteq \mathfrak{i}_{\mu}(A)$. Thus $\mathfrak{i}_{\mu}(A)$ is uncountable and hence $A$ is uncountable.
(ii) Let $x \in c_{\mu}(A)$ and $G$ be an $\omega_{\mu}^{*}$-open set containing $x$. Then there exist $\mathrm{U}_{x} \in \mu$ containing $x$ and a countable subset $C$ such that $U_{x} \backslash C \subseteq \mathfrak{i}_{\mu}(G)$. Then $\left(\mathrm{U}_{x} \backslash \mathrm{C}\right) \cap A \subseteq \mathfrak{i}_{\mu}(\mathrm{G}) \cap A$ i.e., $\left(\mathrm{U}_{x} \cap A\right) \backslash C \subseteq \mathfrak{i}_{\mu}(\mathrm{G}) \cap A$. Since $\mathrm{U}_{x}$ is a $\mu$-open set, $\mathrm{U}_{x} \cap A \neq \varnothing$ and thus $\mathrm{U}_{x} \cap A$ is a non-empty $\mu$-open set. Hence by anti $\mu$-locally countableness of $(X, \mu)$, it follows that $U_{x} \cap A$ is uncountable and hence $\left(\mathrm{U}_{x} \cap A\right) \backslash C$ is also uncountable. Thus $i_{\mu}(G) \cap A$ is uncountable. Hence $\mathrm{G} \cap A \neq \varnothing$. So $x \in \mathrm{c}_{\omega_{\mu}^{*}}(A)$ i.e, $\mathrm{c}_{\mu}(A) \subseteq \mathrm{c}_{\omega_{\mu}^{*}}(A)$. The other part is obvious.

Definition $4 A \mu$-space $(\mathrm{X}, \mu)$ is said to be $\mu$-Lindelöf $[17]$ if every cover of $X$ by $\mu$-open sets has a countable subcover.

A subset A of a $\mu$-space $(\mathrm{X}, \mu)$ is said to be $\mu$-Lindelöf relative to X [17] if every cover of A by $\mu$-open sets of X has a countable subcover.

Theorem 7 Let $(X, \mu)$ be a $\mu$-Lindelöf GTS and $\mathcal{A}$ be a $\mu$-closed, $\omega_{\mu}^{*}$-open subset of $X$. Then $A \backslash \mathfrak{i}_{\mu}(\mathcal{A})$ is countable.

Proof. Clearly $\mathcal{A}$ is a $\mu$-Lindelöf space (see Corollary 3.6 of [15]). For each $x \in A$, there exist $\mathrm{U}_{x} \in \mu$ containing $x$ and a countable subset $C$ such that $\mathrm{U}_{x} \backslash C \subseteq \mathfrak{i}_{\mu}(A)$. Thus $\left\{\mathrm{U}_{x}: x \in A\right\}$ is cover of $A$ by $\mu$-open subsets of $X$. Hence by $\mu$-Lindelöfness of $A$, it has a countable subcover $\left\{U_{n}: n \in \mathbb{N}\right\}$. Since $A \backslash \mathfrak{i}_{\mu}(A) \subseteq \cup\left\{U_{n} \backslash \mathfrak{i}_{\mu}(A): n \in \mathbb{N}\right\}, A \backslash i_{\mu}(A)$ becomes countable.

## 3 Decomposition of continuity by $\omega_{\mu}^{*}$-open sets

Definition 5 A function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ is said to be $\omega_{\mu}^{*}$-continuous (resp. $\omega_{\mu}$-continuous [15], $(\mu, \lambda)$-continuous [1]) if for every $x \in X$ and every $\lambda$-open set V of Y containing $f(\mathrm{x})$, there exists an $\omega_{\mu}^{*}$-open (resp. $\omega_{\mu}$-open, $\mu$-open) set U containing $\chi$ such that $\mathrm{f}(\mathrm{U}) \cong \mathrm{V}$.

Definition 6 A function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ is said to be weakly $\omega_{\mu}^{*}$-continuous if for every $\mathrm{x} \in \mathrm{X}$ and every $\lambda$-open set V of Y containing $\mathrm{f}(\mathrm{x})$, there exists an $\omega_{\mu}^{*}$-open set U containing x such that $\mathrm{f}(\mathrm{U}) \cong \mathrm{c}_{\lambda}(\mathrm{V})$.

Theorem 8 For a function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ the following properties are equivalent:
(i) fis $\omega_{\mu}^{*}$-continuous;
(ii) $\mathrm{f}:\left(\mathrm{X}, \omega_{\mu}^{*}\right) \rightarrow(\mathrm{Y}, \lambda)$ is $\left(\omega_{\mu}^{*}, \lambda\right)$-continuous;
(iii) $\mathrm{f}^{-1}(\mathrm{~V}) \in \omega_{\mu}^{*}$ for every $\mathrm{V} \in \lambda$.

Proof. Obvious.
Remark 5 Let $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ be a function. Then the following relations hold: $(\mu, \lambda)$-continuity $\Rightarrow \omega_{\mu}^{*}$-continuity $\Rightarrow \omega_{\mu}$-continuity $\Rightarrow$ weakly $\omega_{\mu}^{*}$ continuity.

Example 3 (a) Let $X=\{a, b, c\}, \mu=\{\varnothing,\{b\},\{a, c\},\{b, c\}, X\}$ and $\lambda=\{\varnothing,\{a, c\}$, $\{a, b\}$,
$X\}$. Then $\mu$ and $\lambda$ are two GT's on X . It can be verified that the identity mapping $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{X}, \lambda)$ is $\omega_{\mu}^{*}$-continuous but not $(\mu, \lambda)$-continuous.
(b) Let $\mathrm{X}=\mathbb{R}, \mu=$ the usual topology on $\mathbb{R}, \mathrm{Y}=\{\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d}\}$ and $\lambda=$ $\{\varnothing, Y,\{a\},\{a, b\},\{c, d\},\{a, b, c\}, Y\}$. Consider the mapping $f:(X, \mu) \rightarrow(Y, \lambda)$ defined by

$$
f(x)=\left\{\begin{array}{lll}
a, & \text { if } & x \in \mathbb{I} \\
b, & \text { if } & x \notin \mathbb{I}
\end{array}\right.
$$

It can be checked that f is $\omega_{\mu}$-continuous but not $\omega_{\mu}^{*}$-continuous.
(c) Let $X=\mathbb{R}$ be the set of real numbers, $\mu=\{\varnothing, \mathbb{R}, \mathbb{I}\}, Y=\{a, b, c, d\}$ and $\lambda=\{\varnothing, Y,\{d\},\{c, d\},\{a, b, c\}\}$. Consider the mapping $f:(X, \mu) \rightarrow(Y, \lambda)$ defined by

$$
f(x)=\left\{\begin{array}{lll}
a, & \text { if } & x \in \mathbb{I} \cup\{0\} \\
b, & \text { if } & x \notin \mathbb{I} \cup\{0\}
\end{array}\right.
$$

It can be verified that f is $\omega_{\mu}$-continuous but not weakly $\omega_{\mu}^{*}$-continuous.
Definition 7 A function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ is said to be $\left(\omega_{\mu}^{*}, \omega\right)$-continuous (resp. $\left(\omega_{\mu}^{*}, \mu\right)$-continuous) if for every $\lambda$-open set $\mathcal{A}$ of $\mathrm{Y}, \mathrm{f}^{-1}(\mathcal{A})$ is an $\left(\omega_{\mu}, \omega\right)$ set (resp. an $\left(\omega_{\mu}, \mu\right)$-set).

Remark 6 Every $(\mu, \lambda)$-continuous function is $\left(\omega_{\mu}^{*}, \mu\right)$-continuous and every $\omega_{\mu}^{*}$-continuous function is $\left(\omega_{\mu}^{*}, \omega\right)$-continuous but the converses are not true.

Example 4 (a) Let $X=\mathbb{R}, \mu=\{\varnothing,(2,3), X\}$. Then $\mu$ is a GT on $X$. Let $B=\left(1, \frac{3}{2}\right)$ and $\lambda=\{\varnothing, B, X \backslash B, X\}$. Consider the mapping $f:(X, \mu) \rightarrow(X, \lambda)$ defined by

$$
f(x)=\left\{\begin{array}{lll}
\frac{5}{4}, & \text { if } & x \in(1,2) \\
4, & \text { if } & x \notin(1,2)
\end{array}\right.
$$

It can be verified that f is $\left(\omega_{\mu}^{*}, \mu\right)$-continuous but not $(\mu, \lambda)$-continuous.
(b) Let $X=\mathbb{R}, \mu=\{\varnothing, \mathbb{I}, X\}$ where $\mathbb{I}$ is the set of irrationals. Then $\mu$ is a GT on X . Consider the mapping $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{X}, \mu)$ defined by

$$
f(x)=\left\{\begin{array}{lll}
\sqrt{2}, & \text { if } & x \in \mathbb{N} \\
1, & \text { if } & x \notin \mathbb{N}
\end{array}\right.
$$

It can be verified that f is $\left(\omega_{\mu}^{*}, \omega\right)$-continuous but not $\omega_{\mu}^{*}$-continuous.
Definition 8 For any subset $\mathcal{A}$ of a $G T S(X, \mu)$, the $\mu$-frontier [16] of $\mathcal{A}$ is denoted by $\mathrm{Fr}_{\mu}(A)$ and defined by $\mathrm{Fr}_{\mu}(A)=\mathrm{c}_{\mu}(A) \cap \mathrm{c}_{\mu}(X \backslash A)$.

Definition 9 A function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \mathrm{\lambda})$ is said to be co-weakly $\left(\omega_{\mu}^{*}, \lambda\right)$ continuous if $\mathrm{f}^{-1}\left(\mathrm{Fr}_{\lambda}(\mathrm{V})\right)$ is $\omega_{\mu}^{*}$-closed in X for every $\lambda$-open set V in Y .

Theorem 9 For a function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ the following are equivalent:
(i) $f$ is $\omega_{\mu}^{*}$-continuous.
(ii) $f$ is $\omega_{\mu}$-continuous and $\left(\omega_{\mu}, \omega\right)$-continuous.

Proof. (i) $\Leftrightarrow$ (ii) Follows from Theorem 3.

Theorem 10 Let $(\mathrm{X}, \mu)$ be a QTS. Then for a function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ the following are equivalent:
(i) $f$ is $\omega_{\mu}^{*}$-continuous.
(ii) f is co-weakly $\left(\omega_{\mu}^{*}, \lambda\right)$-continuous and weakly $\omega_{\mu}^{*}$-continuous.

Proof. (i) $\Rightarrow$ (ii): Obvious.
(ii) $\Rightarrow$ (i): Let f be a co-weakly $\left(\omega_{\mu}^{*}, \lambda\right)$-continuous and weakly $\omega_{\mu}^{*}$-continuous function. Let $x \in X$ and $V$ be a $\lambda$-open set of $Y$ containing $f(x)$. As $f$ is a weakly $\omega_{\mu}^{*}$-continuous function, there exists a $\omega_{\mu}^{*}$-open set $U$ containing $x$ such that $f(U) \subseteq c_{\lambda}(V)$. Since $F r_{\lambda}(V)=c_{\lambda}(V) \cap c_{\lambda}(X \backslash V)=c_{\lambda}(V) \backslash V$, we have $f(x) \notin \operatorname{Fr}_{\lambda}(V)$. Since $f$ is co-weakly ( $\left.\omega_{\mu}^{*}, \lambda\right)$-continuous, $x \in U \backslash f^{-1}\left(\operatorname{Fr}_{\lambda}(V)\right)$, which is $\omega_{\mu}^{*}$-open in $X$. Then for every $y \in f\left(U \backslash f^{-1}\left(\operatorname{Fr}_{\lambda}(V)\right)\right), y=f\left(x_{1}\right)$ for a point $x_{1} \in U \backslash f^{-1}\left(\operatorname{Fr}_{\lambda}(V)\right)$. Thus we have $f\left(x_{1}\right)=y \in f(U) \subseteq c_{\lambda}(V)$ and $y \notin \operatorname{Fr}_{\lambda}(V)$ with $f\left(x_{1}\right) \in V$. Thus $f\left(U \backslash f^{-1}\left(\operatorname{Fr}_{\lambda}(V)\right)\right) \subseteq V$. Hence $f$ is $\omega_{\mu}^{*}$-continuous.

Theorem 11 For a function $\mathrm{f}:(\mathrm{X}, \mu) \rightarrow(\mathrm{Y}, \lambda)$ the following are equivalent:
(i) $f(\mu, \lambda)$-continuous.
(ii) f is $\omega_{\mu}^{*}$-continuous and $\left(\omega_{\mu}^{*}, \mu\right)$-continuous.

Proof. Follows from Theorem 4.

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## References

[1] Á. Császár, Generalized topology, generalized continuity, Acta Math. Hungar., 96 (2002), 351-357.
[2] Á. Császár, Generalized open sets in generalized topologies, Acta Math. Hungar., 106 (2005), 53-66.
[3] Á. Császár, Remarks on quasi topologies, Acta Math. Hungar., 119 (2008), 197-200.
[4] Á. Császár, On generalized neighbourhood systems, Acta Math. Hungar., 121 (2008), 359-400.
[5] J. Dontchev and M. Przemski, On various decompositions of continuous and some weakly continuous functions, Acta Math. Hungar., 71 (1996), 109-120.
[6] Y. Erguang and Y. Pengfei, On decomposition of $\mathcal{A}$ continuity, Acta Math. Hungar., 110 (2006), 309-313.
[7] M. Ganster and I. L. Reilly, A decomposition of continuity, Acta Math. Hungar., 56 (1990), 299-301.
[8] M. Ganster and I. L. Reilly, Locally closed sets and LC continuous functions, Internat. J. Math. Math. Sci., 12 (1989), 417-424.
[9] A. Gupta and R. D. Sarma, PS-regular sets in topology and generalized topology, Journal of Mathematics, (2014), Article ID 274592, 6 pages.
[10] E. Hatir and T. Noiri, Decomposition of continuity and complete continuity, Acta Math. Hungar., 113 (2006), 281-287.
[11] E. Hatir, T. Noiri and S. Yuksel, A decomposition of continuity, Acta Math Hungar., 70 (1996), 145-150.
[12] H. Z. Hdeib, $\omega$-continuous functions, Dirasat Jour., 16, (2)(1989), 136153.
[13] T. Noiri and O. R. Sayed, On decomposition of continuity, Acta Math. Hungar., 110 (2006), 307-314.
[14] M. Przemski, A decomposition of continuity and $\alpha$-continuity, Acta Math. Hungar., 61 (1993), 93-98.
[15] B. Roy, More on $\mu$-Lindelöf spaces in $\mu$-spaces, Questions and Answers in Gen. Topol., 33 (2015), 25-31.
[16] B. Roy, On weakly ( $\mu, \lambda$ )-open functions, Ukrainian Math Jour., 66 (10) (2015), 1595-1602.
[17] M. S. Sarsak, On $\mu$-compact sets in $\mu$-spaces, Questions and Answers in Gen. Topol., 31 (1) (2013) 49-57.
[18] J. Tong, On decomposition of continuity in topological spaces, Acta Math. Hungar., 54 (1989), 51-55.
[19] J. Tong, Classification of weak continuities and decomposition of continuity, Internat. J. Math. Math. Sci., 51 (2004), 2755-2760.
[20] T. H. Yalvac, Decompositions of continuity, Acta Math. Hungar., 64 (1994), 309-313.

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