# Fuzzy Lie ideals of Lie algebras with interval-valued membership functions 

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#### Abstract

The concept of interval-valued fuzzy sets was first introduced by Zadeh in 1975 as a generalization of fuzzy sets. In this paper we introduce the notion of interval-valued fuzzy Lie ideals in Lie algebras and investigate some of their properties. Using interval-valued fuzzy Lie ideals, characterizations of Noetherian Lie algebras are established. Construction of a quotient Lie algebra via interval-valued fuzzy Lie ideal in a Lie algebra is given. The interval-valued fuzzy isomorphism theorems are also established.


## 1. Introduction

Lie algebras were discovered by Sophus Lie (1842-1899) when he first attempted to classify certain "smooth" subgroups of general linear groups. The groups he considered are now called Lie groups. By taking the tangent space at the identity element of such a group, he obtained the Lie algebra and hence the problems on groups can be reduced to problems on Lie algebras so that it becomes more tractable. Lie algebra is applied in different domains such as physics, hyperbolic and stochastic differential equations. Lie algebra is also largely used by electrical engineers, mainly in the mobile robot control [5].

The notion of interval-valued fuzzy sets was first introduced by Zadeh [13] in 1975 as a generalization of fuzzy sets. An interval-valued fuzzy set is a fuzzy set whose membership function is many-valued and forms an interval with respect to the membership scale. This idea gives the simplest method

[^0]to capture the imprecision of the membership grades for a fuzzy set. It has been noted by Atansassov [3] that such fuzzy sets have some applications in the technological scheme of the functioning of a silo-farm with pneumatic transportation in a plastic products company and in medicine. The interval valued fuzzy sets provide a more adequate description of uncertainty than the traditional fuzzy sets, it is therefore important to use interval valued fuzzy sets in applications. One of the main applications of fuzzy sets is fuzzy control, and one of the most computationally intensive part of fuzzy control is the defuzzification. Since a transition to interval valued fuzzy sets usually increase the amount of computations, it is vitally important to design faster algorithms for the corresponding defuzzification. Fuzzy and anti fuzzy Lie ideals in Lie algebras have been studied in $[1,6,7,10,11,12]$. In this paper, we apply the concept of interval-valued fuzzy sets to Lie algebras. We introduce the notion of interval-valued fuzzy Lie ideals in Lie algebras and investigate some of their properties. Using interval-valued fuzzy Lie ideals, characterizations of Noetherian Lie algebras are established. Construction of a quotient Lie algebra via interval-valued fuzzy Lie ideal in a Lie algebra is given. The interval-valued fuzzy isomorphism theorems are also established.

## 2. Preliminaries

In this paper by $L$ will be denoted a Lie algebra, i.e., a vector space $L$ over a field $F$ (equal to $\mathbf{R}$ or $\mathbf{C}$ ) on which the operation $L \times L \rightarrow L$ denoted by $(x, y) \rightarrow[x, y]$ is defined and satisfies the following axioms:
$\left(L_{1}\right)[x, y]$ is bilinear,
$\left(L_{2}\right) \quad[x, x]=0$ for all $x \in L$,
( $L_{3}$ ) $[[x, y], z]+[[y, z], x]+[[z, x], y]=0$ for all $x, y, z \in L$.
A subspace $H$ of $L$ closed under [, ] will be called a Lie subalgebra. A subspace $I$ of $L$ with the property $[I, L] \subseteq I$ will be called a Lie ideal of $L$. Obviously, any Lie ideal is a subalgebra.

A fuzzy set $\mu: L \rightarrow[0,1]$ is called a fuzzy Lie subalgebra of $L$ if
(a) $\mu(x+y) \geqslant \min \{\mu(x), \mu(y)\}$,
(b) $\mu(\alpha x) \geqslant \mu(x)$,
(c) $\mu([x, y]) \geqslant \min \{\mu(x), \mu(y)\}$
hold for all $x, y \in L$ and $\alpha \in F$.
According to [1] a fuzzy subset $\mu: L \rightarrow[0,1]$ satisfying $(a),(b)$ and
(d) $\mu([x, y]) \geqslant \mu(x)$
is called a fuzzy Lie ideal of $L$. A fuzzy ideal of $L$ is a fuzzy subalgebra [6] such that $\mu(-x) \geqslant \mu(x)$ holds for all $x \in L$.

By an interval number $D$ we mean an interval $\left[a^{-}, a^{+}\right]$, where $0 \leqslant a^{-} \leqslant$ $a^{+} \leqslant 1$. The set of all interval numbers is denoted by $\mathcal{D}[0,1]$. For interval numbers $D_{1}=\left[a_{1}^{-}, b_{1}^{+}\right]$, $D_{2}=\left[a_{2}^{-}, b_{2}^{+}\right]$, we define

$$
\begin{aligned}
& \min \left(D_{1}, D_{2}\right)=\min \left(\left[a_{1}^{-}, b_{1}^{+}\right],\left[a_{2}^{-}, b_{2}^{+}\right]\right)=\left[\min \left\{a_{1}^{-}, a_{2}^{-}\right\}, \min \left\{b_{1}^{+}, b_{2}^{+}\right\}\right], \\
& D_{1} \leqslant D_{2} \longleftrightarrow a_{1}^{-} \leqslant a_{2}^{-} \text {and } b_{1}^{+} \leqslant b_{2}^{+}, \\
& D_{1}=D_{2} \longleftrightarrow a_{1}^{-}=a_{2}^{-} \text {and } b_{1}^{+}=b_{2}^{+} .
\end{aligned}
$$

An interval-valued fuzzy set (briefly, IF set) $A$ on $L$ is defined by

$$
A=\left\{\left(x,\left[\mu_{A}^{-}, \mu_{A}^{+}\right]\right): x \in L\right\}
$$

where $\mu_{A}^{-}$and $\mu_{A}^{+}$are fuzzy sets of $L$ such that $\mu_{A}^{-}(x) \leqslant \mu_{A}^{+}(x)$ for all $x \in L$. Let $\widetilde{\mu}_{A}(x)=\left[\mu_{A}^{-}(x), \mu_{A}^{+}(x)\right]$, then

$$
A=\left\{\left(x, \widetilde{\mu}_{A}(x)\right): x \in L\right\}
$$

where $\widetilde{\mu}_{A}: L \rightarrow \mathcal{D}[0,1]$. For $[s, t] \in \mathcal{D}[0,1]$, the set

$$
U(\widetilde{\mu} ;[s, t])=\{x \in L: \widetilde{\mu}(x) \geq[s, t]\}
$$

is called upper level of $\widetilde{\mu}$.

## 3. Interval-valued fuzzy Lie ideals in Lie algebras

Definition 3.1. An interval-valued fuzzy set $\widetilde{\mu}$ in a Lie algebra $L$ is called an interval-valued fuzzy Lie subalgebra of $L$ if
(1) $\widetilde{\mu}(x+y) \geqslant \min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\}$,
(2) $\widetilde{\mu}(\alpha x) \geqslant \widetilde{\mu}(x)$,
(3) $\widetilde{\mu}([x, y]) \geqslant \min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\}$
hold for all $x, y \in L$ and $\alpha \in F$.

Definition 3.2. An interval-valued fuzzy set $\widetilde{\mu}$ satisfying (1), (2) and
(4) $\widetilde{\mu}([x, y]) \geqslant \widetilde{\mu}(x)$
is called an interval-valued fuzzy Lie ideal of $L$.
From (2) it follows that
(5) $\widetilde{\mu}(0) \geqslant \widetilde{\mu}(x)$,
(6) $\widetilde{\mu}(-x) \geqslant \widetilde{\mu}(x)$
for all $x \in L$.
Example 3.3. The set $\Re^{3}=\{(x, y, z): x, y, z \in R\}$ with the operation $[x, y]=x \times y$, is a real Lie algebra. We define an IF set $\widetilde{\mu}: \Re^{3} \rightarrow \mathcal{D}[0,1]$ by

$$
\widetilde{\mu}(x, y, z)= \begin{cases}{\left[s_{1}, s_{2}\right]} & \text { if } x=y=z=0 \\ {\left[t_{1}, t_{2}\right]} & \text { otherwise }\end{cases}
$$

where $\left[s_{1}, s_{2}\right]>\left[t_{1}, t_{2}\right]$ and $\left[s_{1}, s_{2}\right],\left[t_{1}, t_{2}\right] \in \mathcal{D}[0,1]$. By routine computations, we can see that it is an IF Lie subalgebra and Lie ideal of $\Re^{3}$.

Proposition 3.4. Every interval-valued fuzzy Lie ideal is an interval-valued fuzzy Lie subalgebra.

The converse of Proposition 3.4 is not true in general.
Example 3.5. Let $\Re^{3}$ and [, ] be as in the previous example. Putting

$$
\widetilde{\mu}(x, y, z)= \begin{cases}{[1,1]} & \text { if } x=y=z=0 \\ {[0.5,0.5]} & \text { if } x \neq 0, y=z=0 \\ {[0,0]} & \text { otherwise }\end{cases}
$$

we obtain an example of an interval-valued fuzzy Lie subalgebra which is not an IF Lie ideal. Indeed,

$$
\begin{aligned}
\widetilde{\mu}([(1,0,0)(1,1,1)]) & =\widetilde{\mu}(0,-1,1)=[0,0], \\
\widetilde{\mu}(1,0,0) & =[0.5,0.5] .
\end{aligned}
$$

That is,

$$
\widetilde{\mu}([(1,0,0)(1,1,1)]) \nsupseteq \widetilde{\mu}(1,0,0) .
$$

Theorem 3.6. An interval-valued fuzzy set $\widetilde{\mu}=\left[\mu^{-}, \mu^{+}\right]$in $L$ is an intervalvalued fuzzy Lie ideal if and only if $\mu^{-}$and $\mu^{+}$are fuzzy Lie ideals of $L$.

Proof. Suppose that $\mu^{-}$and $\mu^{+}$are fuzzy Lie ideals of $L$. Then

$$
\begin{aligned}
\widetilde{\mu}(x+y) & =\left[\mu^{-}(x+y), \mu^{+}(x+y)\right] \\
& \left.\left.\geqslant\left[\min \left\{\mu^{-}(x)\right), \mu^{-}(y)\right\}, \min \left\{\mu^{+}(x)\right), \mu^{+}(y)\right\}\right] \\
& =\left[\min \left\{\mu^{-}(x), \mu^{+}(x)\right\}, \min \left\{\mu^{-}(y), \mu^{+}(y)\right\}\right] \\
& =\min \{\widetilde{\mu}(x)), \widetilde{\mu}(y)\}
\end{aligned}
$$

for $x, y \in L$. The verification of (2) and (4) is analogous. Hence $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L$.

Conversely, assume that $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L$. Then

$$
\begin{aligned}
{\left[\mu^{-}(x+y), \mu^{+}(x+y)\right] } & =\widetilde{\mu}(x+y) \geqslant \min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\} \\
& \left.=\min \left\{\left[\mu^{-}(x)\right), \mu^{+}(x)\right],\left[\mu^{-}(y), \mu^{+}(y)\right]\right\} \\
& =\left[\min \left\{\mu^{-}(x), \mu^{-}(y)\right\}, \min \left\{\mu^{+}(x), \mu^{+}(y)\right\}\right]
\end{aligned}
$$

for $x, y \in L$. So,

$$
\mu^{-}(x+y) \geqslant \min \left\{\mu^{-}(x), \mu^{-}(y)\right\} \text { and } \mu^{+}(x+y) \geqslant \min \left\{\mu^{+}(x), \mu^{+}(y)\right\}
$$

In a similar way we can verify (2) and (4). This means that $\mu^{-}$and $\mu^{+}$are fuzzy Lie ideals of $L$.

Theorem 3.7. All nonempty upper levels of interval-valued Lie ideals of a Lie algebra $L$ are Lie ideals of $L$.

Proof. Assume that $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L$ and let $\left[t_{1}, t_{2}\right] \in \mathcal{D}[0,1]$ be such that $U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right) \neq \emptyset$. If $x \in U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$, and $y \in U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$, then $\widetilde{\mu}(x) \geqslant\left[t_{1}, t_{2}\right]$ and $\widetilde{\mu}(y) \geqslant\left[t_{1}, t_{2}\right]$. Hence

$$
\begin{gathered}
\widetilde{\mu}(x+y) \geqslant \min (\widetilde{\mu}(x), \widetilde{\mu}(y)) \geqslant\left[t_{1}, t_{2}\right] \\
\widetilde{\mu}(\alpha x) \geqslant \widetilde{\mu}(x) \geqslant\left[t_{1}, t_{2}\right] \\
\widetilde{\mu}([x, y]) \geqslant \widetilde{\mu}(x) \geqslant\left[t_{1}, t_{2}\right] .
\end{gathered}
$$

So, $x+y \in U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right), \alpha x \in U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$ and $[x, y] \in U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$. This proves that $U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$ is a Lie ideal of $L$.

Definition 3.8. Let $f: L_{1} \rightarrow L_{2}$ be a homomorphism of Lie algebras. For any interval-valued fuzzy set $\widetilde{\mu}$ in a Lie algebra $L_{2}$, we define an intervalvalued fuzzy set $\widetilde{\mu}^{f}$ in $L$ by $\widetilde{\mu}^{f}(x)=\widetilde{\mu}(f(x))$ for all $x \in L_{1}$.

Lemma 3.9. Let $f: L_{1} \rightarrow L_{2}$ be a homomorphism of Lie algebras. If $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L_{2}$, then $\widetilde{\mu}^{f}$ is an interval-valued fuzzy Lie ideal of $L_{1}$.

Proof. Let $x, y \in L_{1}$ and $\alpha \in F$. Then

$$
\begin{aligned}
\widetilde{\mu}^{f}(x+y)=\widetilde{\mu}(f(x+y)) & =\widetilde{\mu}(f(x)+f(y)) \geqslant \min \{\widetilde{\mu}(f(x)), \widetilde{\mu}(f(y))\} \\
& =\min \left\{\widetilde{\mu}^{f}(x), \widetilde{\mu}^{f}(y)\right\}, \\
\widetilde{\mu}^{f}(\alpha x)=\widetilde{\mu}(f(\alpha x)) & =\widetilde{\mu}(\alpha f(x)) \geqslant \widetilde{\mu}(f(x))=\mu^{f}(x), \\
\widetilde{\mu}^{f}([x, y])=\widetilde{\mu}(f([x, y])) & =\widetilde{\mu}([f(x), f(y)]) \geqslant \widetilde{\mu}(f(x))=\widetilde{\mu}^{f}(x),
\end{aligned}
$$

which proves that $\widetilde{\mu}^{f}$ is an interval-valued fuzzy Lie ideal of $L_{1}$.
Theorem 3.10. Let $f: L_{1} \rightarrow L_{2}$ be an epimorphism of Lie algebras. Then $\widetilde{\mu}^{f}$ is an interval-valued fuzzy Lie ideal of $L_{1}$ if and only if $\widetilde{\mu}$ is an intervalvalued fuzzy Lie ideal of $L_{2}$.

Proof. The sufficiency follows from Lemma 3.9. To prove the necessity observe that $f$ is surjective, so for any $x, y \in L_{2}$ there are $x_{1}, y_{1} \in L_{1}$ such that $x=f\left(x_{1}\right), y=f\left(y_{1}\right)$. Thus $\widetilde{\mu}(x)=\widetilde{\mu}^{f}\left(x_{1}\right), \widetilde{\mu}(y)=\widetilde{\mu}^{f}\left(y_{1}\right)$, whence

$$
\begin{gathered}
\widetilde{\mu}(x+y)=\widetilde{\mu}\left(f\left(x_{1}\right)+f\left(y_{1}\right)\right)=\widetilde{\mu}\left(f\left(x_{1}+y_{1}\right)\right)=\widetilde{\mu}^{f}\left(x_{1}+y_{1}\right) \\
\geqslant \min \left\{\widetilde{\mu}^{f}\left(x_{1}\right), \widetilde{\mu}^{f}\left(y_{1}\right)\right\}=\min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\}, \\
\widetilde{\mu}(\alpha x)=\widetilde{\mu}\left(\alpha f\left(x_{1}\right)\right)=\widetilde{\mu}\left(f\left(\alpha x_{1}\right)\right)=\widetilde{\mu}^{f}\left(\alpha x_{1}\right) \geqslant \widetilde{\mu}^{f}\left(x_{1}\right)=\widetilde{\mu}(x), \\
\widetilde{\mu}([x, y])=\widetilde{\mu}\left(\left[f\left(x_{1}\right), f\left(y_{1}\right)\right]\right)=\widetilde{\mu}\left(f\left(\left[x_{1}, y_{1}\right]\right)\right)=\widetilde{\mu}^{f}\left(\left[x_{1}, y_{1}\right]\right) \geqslant \widetilde{\mu}^{f}\left(x_{1}\right)=\widetilde{\mu}(x) .
\end{gathered}
$$

This proves that $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L_{2}$.
Definition 3.11. Two interval-valued fuzzy ideals $\widetilde{\mu}$ and $\widetilde{\lambda}$ of $L$ have the same type if there exists $f \in \operatorname{Aut}(L)$ such that $\widetilde{\mu}(x)=\widetilde{\lambda}(f(x))$ for all $x \in L$.
Theorem 3.12. Let $\widetilde{\mu}$ and $\widetilde{\lambda}$ be interval-valued fuzzy Lie ideals of $L$. Then the following are equivalent:
(i) $\widetilde{\mu}$ and $\widetilde{\lambda}$ have the same type,
(ii) $\widetilde{\mu} \circ f=\widetilde{\lambda}$ for some $f \in \operatorname{Aut}(L)$,
(iii) $g(\widetilde{\mu})=\widetilde{\lambda}$ for some $g \in \operatorname{Aut}(L)$,
(iv) $h(\widetilde{\lambda})=\widetilde{\mu}$ for some $h \in \operatorname{Aut}(L)$,
(v) there exist $h \in \operatorname{Aut}(L)$ such that $U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)=h\left(U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right)$ for all $\left[t_{1}, t_{2}\right] \in \mathcal{D}[0,1]$.

Proof. $(i) \rightarrow(i i)$ : Proof follows immediately from the definition.
(ii) $\rightarrow$ (iii): Suppose that $\widetilde{\mu} \circ f=\widetilde{\lambda}$ for some $f \in \operatorname{Aut}(L)$. Then $\widetilde{\mu}(f(x))=\widetilde{\lambda}(x)$ and $f^{-1}(\widetilde{\mu})(x)=\sup _{y \in f(x)} \widetilde{\mu}(y)=\widetilde{\mu}(f(x))=\widetilde{\lambda}(x)$ for all $x \in L$. If $g=f^{-1}$, then $g \in A u t(L)$ and $g(\widetilde{\mu})=\widetilde{\lambda}$.
$(i i i) \rightarrow(i v)$ : Suppose that $g(\widetilde{\mu})=\widetilde{\lambda}$ for some $g \in A u t(L)$. Then $\widetilde{\sim}(x)=$ $g(\widetilde{\mu})=\sup _{y \in g^{-1}(x)} \widetilde{\mu}(y)=\widetilde{\mu}\left(g^{-1}(x)\right)$. Hence $g^{-1}(x)=\sup _{y \in g(x)} \widetilde{\lambda}(y)=$ $\widetilde{\lambda}(g(y))=\widetilde{\mu}\left(g^{-1}(g(x))\right)=\widetilde{\mu}(x)$ for all $x \in L$. If $h=g^{-1}$, then $h \in \operatorname{Aut}(L)$ and $h(\widetilde{\lambda})=\widetilde{\mu}$.
$(i v) \rightarrow(v):$ If $h(\widetilde{\lambda})=\widetilde{\mu}$ for some $h \in \operatorname{Aut}(L)$, then $\widetilde{\mu}(x)=h(\widetilde{\lambda})(x)=$ $\sup _{y \in h^{-1}(x)} \widetilde{\lambda}(y)=\widetilde{\lambda}\left(h^{-1}(x)\right)$ for all $x \in L$.

Let $\left[t_{1}, t_{2}\right] \in \mathcal{D}[0,1]$. We need to show $U\left(\tilde{\mu} ;\left[t_{1}, t_{2}\right]\right)=h\left(U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right)$. If $x \in U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$, then $\left.\widetilde{\lambda}\left(h^{-1}(x)\right)=\widetilde{\mu} \geqslant\left[t_{1}, t_{2}\right]\right]$ which implies that $h^{-1}(x) \in U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)$, i.e., $x \in h\left(U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right)$. Thus we obtain $U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$ $\subseteq h\left(\underset{\sim}{U}\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right)$. On the other hand, let $x \in h\left(U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right)$. Then $h^{-1}(x)$ $\in U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)$ and so $\widetilde{\mu}(x)=\widetilde{\lambda}\left(h^{-1}(x)\right) \geqslant\left[t_{1}, t_{2}\right]$. It follows that $x \in$ $U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$. Hence $h\left(U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right) \subseteq U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)$ and $(v)$ holds.
$(v) \rightarrow(i):$ Suppose that there exists $h \in \operatorname{Aut}(L)$ such that $U\left(\widetilde{\mu} ;\left[t_{1}, t_{2}\right]\right)=$ $h\left(U\left(\widetilde{\lambda} ;\left[t_{1}, t_{2}\right]\right)\right)$ for all $\left[t_{1}, t_{2}\right] \in \mathcal{D}[0,1]$. Let $\widetilde{\lambda}\left(h^{-1}(x)\right)=\left[s_{1}, s_{2}\right]$. Then $h^{-1}(x) \in U\left(\lambda ;\left[s_{1}, s_{2}\right]\right)$, hence $x \in h\left(U\left(\widetilde{\lambda} ;\left[s_{1}, s_{2}\right]\right)\right)=U\left(\widetilde{\mu} ;\left[s_{1}, s_{2}\right]\right)$. Thus $\widetilde{\mu}(x) \geqslant\left[s_{1}, s_{2}\right]=\widetilde{\lambda}\left(h^{-1}(x)\right)$. Hence $\widetilde{\mu}(x)=\widetilde{\lambda}\left(h^{-1}(x)\right)$ for all $x \in L$, which proves that $\widetilde{\mu}$ and $\tilde{\lambda}$ have the same type.

## 4. Characterizations of Noetherian Lie algebras

Definition 4.1. A Lie algebra $L$ is said to be Noetherian if for every ascending sequence $I_{1} \subseteq I_{2} \subseteq I_{3} \subseteq \ldots$ of Lie ideals of $L$ there exists a natural number $n$ such that $I_{n}=I_{k}$ for all $n \geqslant k$.

Theorem 4.2. A Lie algebra $L$ is Noetherian if and only if the set of values of any its interval-valued fuzzy Lie ideal is well-ordered.

Proof. Suppose that $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal whose set of values is not well-ordered subset of $\mathcal{D}[0,1]$. Then there exists a strictly decreasing sequence $\left\{\left[\alpha_{n}, \beta_{n}\right]\right\}$ such that $\left[\alpha_{n}, \beta_{n}\right]=\widetilde{\mu}\left(x_{n}\right)$ for some $x_{n} \in L$. Let $B_{n}:=\left\{x \in L \mid \widetilde{\mu}(x) \geqslant\left[\alpha_{n}, \beta_{n}\right]\right\}$. Then $B_{1} \subset B_{2} \subset B_{3} \subset \ldots$ form
a strictly ascending chain of Lie ideals of $L$, contradicting the assumption that $L$ is Noetherian.

Conversely, suppose that the set of values of any interval-valued fuzzy Lie ideal of $L$ but $L$ is not Noetherian. Then there exists a strictly ascending chain $A_{1} \subset A_{2} \subset A_{3} \subset \ldots$ of Lie ideals of $L$. Suppose that $A=\bigcup_{k=1}^{\infty} A_{k}$ is a Lie ideal of $L$. Define an interval-valued fuzzy set $\widetilde{\mu}$ in $L$ by putting

$$
\widetilde{\mu}(x):= \begin{cases}{\left[\frac{1}{k+1}, \frac{1}{k}\right]} & \text { for } x \in A_{k} \backslash A_{k-1}, \\ {[0,0]} & \text { for } x \notin A .\end{cases}
$$

We claim that $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideals of $L$. Let $x, y \in L$.
If $x, y \in A$ then there are $m, n$ such that $x \in A_{n} \backslash A_{n-1}, y \in A_{m} \backslash A_{m-1}$. Obviously $x+y \in A_{k} \backslash A_{k-1} \subset A_{p}$, where $k \leqslant p=\max \{m, n\}$. So, $\widetilde{\mu}(x)=$ $\left[\frac{1}{n+1}, \frac{1}{n}\right], \widetilde{\mu}(y)=\left[\frac{1}{m+1}, \frac{1}{m}\right]$ and

$$
\widetilde{\mu}(x+y)=\left[\frac{1}{k+1}, \frac{1}{k}\right] \geqslant\left[\frac{1}{p+1}, \frac{1}{p}\right]=\min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\} .
$$

In the case $x \notin A, y \in A$ we have $y \in A_{m} \backslash A_{m-1}$ for some natural $m$. Hence $\widetilde{\mu}(x)=[0,0], \widetilde{\mu}(y)=\left[\frac{1}{m+1}, \frac{1}{m}\right]$, consequently

$$
\widetilde{\mu}(x+y) \geqslant[0,0]=\min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\} .
$$

The case $x \in A, y \notin A$ is analogous. The case $x \notin A, y \notin A$ is obvious. The verification of (2) and (4) is analogous. Thus $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L$. Consequently, $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal. Since the chain $A_{1} \subset A_{2} \subset A_{3} \subset \ldots$ is not terminating, $\widetilde{\mu}$ has a strictly descending sequence of values. This contradicts that the value set of any interval-valued fuzzy Lie ideal is well-ordered. This completes the proof.

We note that a set is well ordered if and only if it does not contain any infinite decreasing sequence.

Theorem 4.3. Let $S=\left\{\left[s_{1}, t_{1}\right],\left[s_{2}, t_{2}\right], \ldots\right\} \cup\{[0,0]\}$, where $\left\{\left[s_{n}, t_{n}\right]\right\}$ is a strictly decreasing sequence in $\mathcal{D}[0,1]$. Then a Lie algebra $L$ is Noetherian if and only if for each interval-valued fuzzy Lie ideal $\widetilde{\mu}$ of $L, \operatorname{Im}(\widetilde{\mu}) \subseteq S$ implies that there exists a positive integer $m$ such that $\operatorname{Im}(\widetilde{\mu}) \subseteq\left\{\left[s_{1}, t_{1}\right],\left[s_{2}, t_{2}\right], \ldots\right.$, $\left.\left[s_{m}, t_{m}\right]\right\} \cup\{[0,0]\}$.

Proof. If $L$ is a Noetherian Lie algebra, then $\operatorname{Im}(\widetilde{\mu})$ is a well ordered subset of $\mathcal{D}[0,1]$.

Conversely, if the above condition is satisfied and $L$ is not Noetherian, then there exists a strictly ascending chain $A_{1} \subset A_{2} \subset A_{3} \subset \ldots$ of Lie ideals of $L$. Define an interval-valued fuzzy set $\widetilde{\mu}$ by

$$
\widetilde{\mu}(x):= \begin{cases}{\left[s_{1}, t_{1}\right]} & \text { if } x \in A_{1}, \\ {\left[s_{n}, t_{n}\right]} & \text { if } x \in A_{n} \backslash A_{n-1}, n=2,3,4, \ldots \\ {[0,0]} & \text { if } x \in G \backslash \bigcup_{n=1}^{\infty} A_{n} .\end{cases}
$$

Let $x, y \in L$. If either $x$ or $y$ belongs to $G \backslash \bigcup_{n=1}^{\infty} A_{n}$, then either $\widetilde{\mu}(x)=[0,0]$ or $\widetilde{\mu}(y)=[0,0]$. Thus $\widetilde{\mu}(x+y) \geqslant \min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\}$.

If $x, y \in A_{1}$, then $x \in A_{1}$ and so $\widetilde{\mu}(x+y)=\left[s_{1}, t_{1}\right] \geqslant \min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\}$.
If $x, y \in A_{n} \backslash A_{n-1}$, then $x \in A_{n}$ and $\widetilde{\mu}(x+y) \geq\left[s_{n}, t_{n}\right]=\min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\}$. Assume that $x \in A_{1}$ and $y \in A_{n} \backslash A_{n-1}$ for $n=2,3,4, \ldots$, then $x+y \in A_{n}$ and hence

$$
\widetilde{\mu}(x+y) \geq\left[s_{n}, t_{n}\right]=\min \left\{\left[s_{1}, t_{1}\right],\left[s_{n}, t_{n}\right]\right\}=\min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\} .
$$

Similarly for $x \in A_{n} \backslash A_{n-1}$ and $y \in A_{1}$ for $n=2,3,4, \ldots$, we have

$$
\widetilde{\mu}(x+y) \geq\left[s_{n}, t_{n}\right]=\min \{\widetilde{\mu}(x), \widetilde{\mu}(y)\} .
$$

Hence $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of Lie algebra. This contradicts our assumption. The verification of (2) and (4) is analogous and we omit the details. This completes the proof.

## 5. Quotient Lie algebra via IF Lie ideals

Theorem 5.1. Let I be a Lie ideal of a Lie algebra L. If $\widetilde{\mu}$ is an intervalvalued Lie ideal of $L$, then an interval-valued fuzzy set $\overline{\widetilde{\mu}}$ defined by

$$
\overline{\widetilde{\mu}}(a+I)=\sup _{x \in I} \widetilde{\mu}(a+x)
$$

is an interval-valued Lie ideal of the quotient Lie algebra $L / I$.
Proof. Clearly, $\overline{\widetilde{\mu}}$ is well-defined. Let $x+I, y+I \in L / I$, then

$$
\begin{aligned}
\overline{\widetilde{\mu}}(x+I)+(y+I)) & =\overline{\widetilde{\mu}}_{A}((x+y)+I)=\sup _{z \in I} \widetilde{\mu}((x+y)+z) \\
& =\sup _{z=s+t \in I} \widetilde{\mu}((x+y)+(s+t)) \\
& \geqslant \sup _{s, t \in I} \min \{\widetilde{\mu}(x+s), \widetilde{\mu}(y+t)\} \\
& =\min \left\{\sup _{s \in I} \widetilde{\mu}(x+s), \sup _{t \in I} \widetilde{\mu}(y+t)\right\} \\
& =\min \{\widetilde{\widetilde{\mu}}(x+I), \widetilde{\widetilde{\mu}}(y+I)\},
\end{aligned}
$$

$$
\begin{gathered}
\overline{\widetilde{\mu}}(\alpha(x+I))=\overline{\widetilde{\mu}}(\alpha x+I)=\sup _{z \in I} \widetilde{\mu}(\alpha x+z) \geqslant \sup _{z \in I} \widetilde{\mu}(x+z)=\widetilde{\widetilde{\mu}}(x+I), \\
\overline{\widetilde{\mu}}([x+I, y+I])=\overline{\widetilde{\mu}}([x, y]+I)=\sup _{z \in I} \widetilde{\mu}([x, y]+z) \\
\\
\geqslant \sup _{z \in I} \widetilde{\mu}(x+z)=\overline{\widetilde{\mu}}(x+I)
\end{gathered}
$$

Hence $\overline{\widetilde{\mu}}$ is an interval-valued fuzzy Lie ideal of $L / I$.
Theorem 5.2. Let $f: L_{1} \rightarrow L_{2}$ be a homomorphism of a Lie algebra $L_{1}$ onto a Lie algebra $L_{2}$.
(i) If $\widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L_{1}$, then $f(\widetilde{\mu})$ is an intervalvalued fuzzy Lie ideal of $L_{2}$,
(ii) If $\widetilde{\lambda}$ is an interval-valued fuzzy Lie ideal of $L_{2}$, then $f^{-1}(\widetilde{\lambda})$ is an interval-valued fuzzy Lie ideal of $L_{1}$.

Proof. Straightforward.
For an interval-valued fuzzy Lie ideal $\widetilde{\mu}$ of a Lie algebra $L$ we define a binary relation $\sim$ by putting

$$
x \sim y \longleftrightarrow \widetilde{\mu}(x-y)=\widetilde{\mu}(0)
$$

This relation is a congruence. The set of all its equivalence classes $\widetilde{\mu}[x]$ is denoted by $L / \widetilde{\mu}$. It is a Lie algebra under the following operations:

$$
\widetilde{\mu}[x]+\widetilde{\mu}[y]=\widetilde{\mu}[x+y], \quad \alpha \widetilde{\mu}[x]=\widetilde{\mu}[\alpha x], \quad[\widetilde{\mu}[x], \widetilde{\mu}[y]]=\widetilde{\mu}[[x, y]]
$$

where $x, y \in L, \alpha \in F$.
Theorem 5.3. (First IF isomorphism theorem)
Let $f: L_{1} \rightarrow L_{2}$ be an epimorphism of Lie algebras and let $\widetilde{\mu}$ be an intervalvalued fuzzy Lie ideal of $L_{2}$. Then $L_{1} / f^{-1}(\widetilde{\mu}) \cong L_{2} / \widetilde{\mu}$.

Proof. Define a map $\theta: L_{1} / f^{-1}(\widetilde{\mu}) \rightarrow L_{2} / \widetilde{\mu}$ by $\theta\left(f^{-1}(\widetilde{\mu})[x]\right)=\widetilde{\mu}[f(x)]$.
$\theta$ is well-defined since $f^{-1}(\widetilde{\mu})[x]=f^{-1}(\widetilde{\mu})[y]$ implies $f^{-1}(\widetilde{\mu})(x-y)=$ $f^{-1}(\widetilde{\mu})(0)$. Whence $\widetilde{\mu}(f(x)-f(y))=\widetilde{\mu}(f(0))=\widetilde{\mu}(0)$. Thus $\widetilde{\mu}[f(x)]=$ $\widetilde{\mu}[f(y)]$.
$\theta$ is one to one because $\widetilde{\mu}[f(x)]=\widetilde{\mu}[f(y)]$ gives $\widetilde{\mu}(f(x)-f(y))=\widetilde{\mu}(0)$, i.e., $\widetilde{\mu}(f(x)-f(y))=\widetilde{\mu}(f(0))$, which proves $f^{-1}(\widetilde{\mu})(x-y)=f^{-1}(\widetilde{\mu})(0)$. Thus $f^{-1}(\widetilde{\mu})[x]=f^{-1}(\widetilde{\mu})[y]$.

Since $f$ is an onto, $\theta$ is an onto. Finally, $\theta$ is a homomorphism because

$$
\begin{aligned}
\theta\left(f^{-1}(\widetilde{\mu})[x]+f^{-1}(\widetilde{\mu})[y]\right) & =\theta\left(f^{-1}(\widetilde{\mu})[x+y]\right)=\widetilde{\mu}[f(x+y)]=\widetilde{\mu}[f(x)+f(y)] \\
& =\widetilde{\mu}[f(x)]+\widetilde{\mu}[f(y)]=\theta\left(f^{-1}(\widetilde{\mu})[x]\right)+\theta\left(f^{-1}(\widetilde{\mu})[y]\right), \\
\theta\left(\alpha f^{-1}(\widetilde{\mu})[x]\right)=\theta\left(f^{-1}(\widetilde{\mu})[\alpha x]\right) & =\widetilde{\mu}[f(\alpha x)]=\alpha \widetilde{\mu}[f(x)]=\alpha \theta\left(f^{-1}(\widetilde{\mu})[x]\right), \\
\theta\left(\left[f^{-1}(\widetilde{\mu})[x], f^{-1}(\widetilde{\mu})[y]\right]\right) & =\theta\left(\left[f^{-1}(\widetilde{\mu})[x, y]\right]\right)=\widetilde{\mu}[f([x, y])] \\
& =\widetilde{\mu}[[f(x), f(y)]]=[\widetilde{\mu}[f(x)], \widetilde{\mu}[f(y)]] \\
& =\left[\theta\left(f^{-1}(\widetilde{\mu})[x]\right), \theta\left(f^{-1}(\widetilde{\mu})[y]\right)\right] .
\end{aligned}
$$

Hence $L_{1} / f^{-1}(\widetilde{\mu}) \cong L_{2} / \widetilde{\mu}$.
We state the following IF isomorphism Theorems without proofs.
Theorem 5.4. (Second IF isomorphism theorem)
Let $\widetilde{\mu}$ and $\widetilde{\lambda}$ be two interval-valued fuzzy subsets of the same Lie algebra. If $\widetilde{\mu}$ is a subalgebra and $\widetilde{\lambda}$ is a Lie ideal, then
(i) $\widetilde{\lambda}$ is an interval-valued fuzzy Lie ideal of $\widetilde{\mu}+\widetilde{\lambda}$,
(ii) $\widetilde{\mu} \cap \widetilde{\lambda}$ is an interval-valued fuzzy ideal of $\widetilde{\mu}$,
(iii) $(\widetilde{\mu}+\widetilde{\lambda}) / \lambda \cong \widetilde{\mu} /(\widetilde{\mu} \cap \widetilde{\lambda})$.

Theorem 5.5. (Third IF isomorphism theorem)
Let $\widetilde{\mu}$ and $\widetilde{\lambda}$ be interval-valued fuzzy Lie ideals of the same Lie algebra such that $\widetilde{\mu} \leqslant \widetilde{\lambda}$. Then
(i) $\tilde{\lambda} / \widetilde{\mu}$ is an interval-valued fuzzy Lie ideal of $L / \widetilde{\mu}$,
(ii) $(L / \widetilde{\mu}) /(\widetilde{\lambda} / \widetilde{\mu}) \cong L / \widetilde{\lambda}$.

Theorem 5.6. (IF Zassenhaus lemma)
Let $\widetilde{\mu}$ and $\widetilde{\lambda}$ be interval-valued fuzzy subalgebras of a Lie algebra $L$ and let $\widetilde{\mu}_{1}$ and $\widetilde{\lambda}_{1}$ be an interval-valued fuzzy Lie ideals of $\widetilde{\mu}$ and $\widetilde{\lambda}$ respectively. Then
(a) $\widetilde{\mu}_{1}+\left(\widetilde{\mu} \cap \widetilde{\lambda}_{1}\right)$ is an interval-valued fuzzy Lie ideal of $\widetilde{\mu}_{1}+(\widetilde{\mu} \cap \widetilde{\lambda})$,
(b) $\widetilde{\lambda}_{1}+\left(\widetilde{\mu}_{1} \cap \widetilde{\lambda}\right)$ is an interval-valued fuzzy ideal of $\widetilde{\lambda}_{1}+(\widetilde{\mu} \cap \widetilde{\lambda})$,
(c) $\frac{\widetilde{\mu}_{1}+(\widetilde{\mu} \cap \tilde{\lambda})}{\widetilde{\mu}_{1}+\left(\widetilde{\mu} \cap \tilde{\lambda}_{1}\right)} \simeq \frac{\tilde{\lambda}_{1}+(\widetilde{\mu} \cap \tilde{\lambda})}{\tilde{\lambda}_{1}+\left(\widetilde{\mu}_{1} \cap \tilde{\lambda}\right)}$.

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# Counting loops with the inverse property 

Asif Ali and John Slaney


#### Abstract

The numbers of isomorphism classes of IP loops of order up to 13 have been obtained by exhaustive enumeration, and are presented here along with some basic observations concerning IP loops.


## 1. Introduction

An $I P$ loop is a set $L$ and a binary operation $*$, where $L$ contains an identity $e$ such that $a * e=a=e * a$ for all $a \in L$, and where each $x \in L$ has a two-sided inverse $x^{-1}$ such that for all $y \in L$

$$
x^{-1} *(x * y)=y=(y * x) * x^{-1}
$$

For an account of the properties of IP loops, see Bruck's survey [3]). Clearly every group is an IP loop, but the converse is not the case. Steiner loops are also IP loops, satisfying the extra condition $x^{-1}=x$. IP loops form a very important class, not only in that they represent a strong generalization of both groups and Steiner loops, but also in that the Moufang nucleus (the set of $a \in L$ such that $a[(x y) a]=(a x)(y a)$ for all $x, y \in L)$ of such loops behaves as a nilpotency function for this class. Moreover IP loops are exactly those groupoids whose power sets are the semiassociative relation algebras [7].

The present paper reports the numbers of non-isomorphic IP loops having order up to 13 . Since these were obtained by exhaustive enumeration, they are available for inspection.

[^1]
## 2. History of counting loops

The number of non-isomorphic loops up to order 6 was found by Schönhardt [12] in 1930, but this was not noticed by Albert [1] or Sade [11] who obtained weaker results much later. Dénes and Keedwell [5] present counts of "quasigroups" up to order 6 , but in fact count loops owing to their assumption that each "quasigroup" is isomorphic to a reduced square, which is obviously untrue of quasigroups in general. The loops of order 7 were counted in 1985 by Brant and Mullen [2]. In 2001, "QSCGZ" announced the number of loops of order 8 in an electronic forum [10], and the same value was found independently by Gujerin. For more on the history of counting loops, see McKay et al [9].

## 3. IP loops of small order

The smallest IP loop which is not a group is of order 7:

| $*$ | 1 | 2 | 3 | 4 | 5 | 6 | 7 |  |  |
| ---: | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- | :--- |
|  | 1 | 2 | 3 | 4 | 5 | 6 | 7 |  |  |
| 2 | 2 | 3 | 1 | 6 | 7 | 5 | 4 | $x$ | $x^{-1}$ |
| 3 | 3 | 1 | 2 | 7 | 6 | 4 | 5 |  | 1 |
| 4 | 4 | 7 | 6 | 5 | 1 | 2 | 3 | 3 | 3 |
| 5 | 5 | 6 | 7 | 1 | 4 | 3 | 2 | 4 | 5 |
| 6 | 6 | 4 | 5 | 3 | 2 | 7 | 1 | 5 | 4 |
| 7 | 7 | 5 | 4 | 2 | 3 | 1 | 6 | 6 | 7 |
| 7 |  | 7 | 6 |  |  |  |  |  |  |

This structure has proper subalgebras $\{1,2,3\},\{1,4,5\}$ and $\{1,6,7\}$. Note that the order of these subloops does not divide the order of the loop, marking a significant difference between IP loops and groups.

Note also that the only element which is its own inverse is the identity $e$. This is a general feature of IP loops of odd order, as may be shown by a simple counting argument:

Observation 1. IP loops of odd order have no subloops of even order.
Proof. Let $(L, *)$ be an IP loop and let $(S, \underline{*})$ be a subloop of $(L, *)$ of even order. Clearly, $S$ consists of $e$ and some subset of elements of $L$ along with their inverses. For this subset to be of even cardinality, some element in it other than $e$ must be self-inverse and thus of order 2 . Let $a \in L$ be such an element of order 2. Let $\dagger x$ be defined as $a * x$. Then the operation $\dagger$ is of period 2, because $\dagger \dagger x=a *(a * x)=a^{-1} *(a * x)=x$. Moreover, $\dagger$ has
no fixed points, because if $\dagger x=x$ then $a * x=x$, so $a=e$, contradicting the assumption that $a$ is of order 2 . Hence $\dagger$ partitions $L$ into pairs, so the cardinality of $L$ must be even.

The IP loops of small orders were counted by using a finite domain constraint solver to generate representatives of all isomorphism classes. The solver FINDER [13] has previously been used to generate results concerning the spectra of quasigroup identities [6]. It works by expressing each equation or other defining condition as the set of its ground instances on the domain of $N$ elements, compiling these into constraints and then conducting a backtracking search for solutions to the constraint satisfaction problem using standard techniques such as forward checking and nogood learning [4].

Some symmetries were broken by enforcing conditions such as that $e$ is always the first element. The remaining isomorphic copies were eliminated in a postprocessing phase. The results to order 11 were independently corroborated using the first order theorem prover PROVER9 and its associated propositional satisfiability solver MACE-4 [8]. In the cases of order 12 and order 13, the required searches are too hard for MACE and PROVER9, so we have only the results by FINDER in those cases.

| size | groups | non - groups | total |
| :---: | :---: | :---: | :---: |
| 1 | 1 | 0 | 1 |
| 2 | 1 | 0 | 1 |
| 3 | 1 | 0 | 1 |
| 4 | 2 | 0 | 2 |
| 5 | 1 | 0 | 1 |
| 6 | 2 | 0 | 2 |
| 7 | 1 | 1 | 2 |
| 8 | 5 | 3 | 8 |
| 9 | 2 | 5 | 7 |
| 10 | 2 | 45 | 47 |
| 11 | 1 | 48 | 49 |
| 12 | 5 | 2679 | 2684 |
| 13 | 1 | 10341 | 10342 |

Table 1. Numbers of IP loops of given order
The full list of these small IP loops, in a simple matrix format as for the order 7 example above, is available online. ${ }^{1}$

[^2]
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# On middle translations of finite quasigroups 

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#### Abstract

We prove that a finite quasigroup is isotopic to a group if and only if some set of bijections induced by middle transformations of this quasigroup is a group.


## 1. Introduction

Let $Q=\{1,2,3, \ldots, n\}$ be a finite set, $\varphi$ and $\psi$ permutations of $Q$. The multiplication (composition) of permutations is defined as $\varphi \psi(x)=\varphi(\psi(x))$.

Let $Q(\cdot)$ be a quasigroup. Permutations $L_{a}: x \rightarrow a \cdot x, R_{a}: x \rightarrow x \cdot a$ are called left and right translations of $Q(\cdot)$. Permutations $\lambda_{i}, \varphi_{i}(i \in Q)$ of $Q$ such that

$$
\begin{align*}
& \lambda_{i}(x) \cdot x=i,  \tag{1}\\
& x \cdot \varphi_{i}(x)=i \tag{2}
\end{align*}
$$

for all $x \in Q$, are called left (respectively: right) middle translations of an element $i$ in a quasigroup $Q(\cdot)$. Such translation were firstly studied by V. D. Belousov (cf. [1]) in connection with some groups associated with quasigroups. Next, the investigations of such translations were continued by many authors, see for example [3] or [5].

The above two conditions say that in a Latin square $n \times n$ connected with a quasigroup $Q(\cdot)$ of order $n$ we select $n$ cells, one in each row, one in each column, containing the same fixed element $i . \lambda_{i}(x)$ means that to find in the column $x$ the cell containing an element $i$ we must select the row $\lambda_{i}(x)$. Analogously, $\varphi_{i}(x)$ means that to find in the row $x$ the cell containing $i$ we must select the column $\varphi_{i}(x)$. Thus, $\lambda_{i}$ is a selection of

[^3]rows, $\varphi_{i}$ - a selection of columns, containing an element $i$. In connection with this fact $\lambda_{i}$ will be called a left track ( $l$-track), $\varphi_{i}$ - a right track ( $r$-track) of an element $i$. It is clear that for a quasigroup $Q(\cdot)$ of order $n$ the set $\left\{\lambda_{1}, \lambda_{2}, \ldots, \lambda_{n}\right\}$ uniquely determines its Latin square, and conversely, any Latin square $n \times n$ uniquely determines the set $\left\{\lambda_{1}, \lambda_{2}, \ldots, \lambda_{n}\right\}$. A similar situation holds for $\left\{\varphi_{1}, \varphi_{2}, \ldots, \varphi_{n}\right\}$.

More interesting facts on connections of translations with Latin squares one can find in [2].

As a simple consequence of the above definitions we obtain
Proposition 1.1. In any quasigroup $Q(\cdot)$ the following identities hold:

1) $\lambda_{i}=\varphi_{i}^{-1}$,
2) $\varphi_{i}^{-1}(x) \cdot x=i$,
3) $L_{i}(x)=\left(\lambda_{i}(x) \cdot x\right) \cdot x$,
4) $\quad L_{i}(x)=\left(x \cdot \varphi_{i}(x)\right) \cdot x$,
5) $\quad R_{i}(x)=x \cdot\left(\lambda_{i}(x) \cdot x\right)$,
6) $\quad R_{i}(x)=x \cdot\left(x \cdot \varphi_{i}(x)\right)$.

Corollary 1.2. In any group $G(\cdot)$ we have

1) $\varphi_{i}(x)=x^{-1} \cdot i, \quad \lambda_{i}(x)=i \cdot x^{-1}$,
2) $\varphi_{1}(x)=\lambda_{1}(x)=x^{-1}$,
3) $L_{i}(x)=\lambda_{i}(x) \cdot x^{2}$,
4) $R_{i}(x)=x^{2} \cdot \varphi_{i}(x)$,
where 1 is the identity element of the group $G(\cdot)$.

## 2. Isotopy invariants in quasigroups

Two quasigroups $Q(\cdot)$ and $Q(\circ)$ are isotopic if there exists an ordered triple $T=(\alpha, \beta, \gamma)$ of bijections $\alpha, \beta, \gamma: Q \rightarrow Q$ such that

$$
\gamma(x \circ y)=\alpha(x) \cdot \beta(y)
$$

for all $x, y \in Q$.
For $y=\psi_{i}(x)$, where $\psi_{i}$ is a $r$-track of a quasigroup $Q(\circ)$, this identity has the form

$$
\gamma\left(x \circ \psi_{i}(x)\right)=\alpha(x) \cdot \beta \psi_{i}(x),
$$

whence, according to (2), we obtain

$$
\gamma(i)=\alpha(x) \cdot \beta \psi_{i}(x)
$$

This for $z=\alpha(x)$ and $j=\gamma(i)$ gives

$$
j=z \cdot \beta \psi_{i} \alpha^{-1}(z) .
$$

Since

$$
j=z \cdot \varphi_{j}(z)=z \cdot \varphi_{\gamma(i)}(z)
$$

for $r$-tracks $\varphi_{j}$ and $\varphi_{\gamma(i)}$ of a quasigroup $Q(\cdot)$, the above implies

$$
\begin{equation*}
\varphi_{\gamma(i)}=\beta \psi_{i} \alpha^{-1} . \tag{3}
\end{equation*}
$$

Remark 2.1. For $l$-tracks $\lambda_{i}$ and $\mu_{i}$ of isotopic quasigroups $Q(\cdot)$ and $Q(\circ)$ we have

$$
\begin{equation*}
\lambda_{\gamma(i)}=\alpha \mu_{i} \beta^{-1} \tag{4}
\end{equation*}
$$

Definition 2.2. By a spin of a quasigroup $Q(\cdot)$ we mean the permutation

$$
\varphi_{i j}=\varphi_{i} \varphi_{j}^{-1}=\varphi_{i} \lambda_{j}
$$

where $\varphi_{i}$ and $\lambda_{j}$ are tracks of $Q(\cdot)$. The spin $\varphi_{i i}$ is called trivial.
The set of all spins of a quasigroup $Q(\cdot)$ is denoted by $\Phi_{Q}(\cdot)$.
Proposition 2.3. Spins have the following properties

1) $\varphi_{i j}(x) \neq x$ for all $x \in Q$ and $i \neq j$,
2) $\varphi_{p i}(x) \neq \varphi_{p j}(x)$ for all $x \in Q$ and $i \neq j$,
3) $\varphi_{i j}=\varphi_{j i}^{-1}$,
4) $\varphi_{k i} \varphi_{i l}=\varphi_{k l}$,
5) $\varphi_{m k}=\varphi_{i m}^{-1} \varphi_{i k}$.

Proof. (1) If $\varphi_{i j}(x)=x$ holds for some $i \neq j$ and $x \in Q$, then, according to the definition of $\varphi_{i j}$, we have $\varphi_{i} \varphi_{j}^{-1}(x)=x$. Whence, for $x=\varphi_{j}(y)$, we obtain $\varphi_{i}(y)=\varphi_{j}(y)$. Consequently $y \cdot \varphi_{i}(y)=y \cdot \varphi_{j}(y)$, i.e., $i=j$. This contradicts our assumption. So, $\varphi_{i j}(x) \neq x$ for all $x \in Q$ and $i \neq j$.
(2) Analogously as (1).
(3) $\varphi_{i j}=\varphi_{i} \varphi_{j}^{-1}=\left(\varphi_{j} \varphi_{i}^{-1}\right)^{-1}=\varphi_{j i}^{-1}$.
(4) $\varphi_{k i} \varphi_{i l}=\left(\varphi_{k} \varphi_{i}^{-1}\right)\left(\varphi_{i} \varphi_{l}^{-1}\right)=\varphi_{k}\left(\varphi_{i}^{-1} \varphi_{i}\right) \varphi_{l}^{-1}=\varphi_{k l}$.
(5) $\varphi_{m k}=\varphi_{m} \varphi_{k}^{-1}=\varphi_{m} \varphi_{i}^{-1} \varphi_{i} \varphi_{k}^{-1}=\left(\varphi_{i} \varphi_{m}^{-1}\right)^{-1}\left(\varphi_{i} \varphi_{k}^{-1}\right)=\varphi_{i m}^{-1} \varphi_{i k}$.

As it is well-known any permutation $\varphi$ of the set $Q$ of order $n$ can be decomposed into $r \leqslant n$ cycles of the length $k_{1}, \ldots, k_{r}$ and $k_{1}+\ldots+k_{r}=n$. We denote this fact by

$$
Z(\varphi)=\left[k_{1}, k_{2}, \ldots, k_{r}\right] .
$$

Since conjugate permutations are decomposable into cycles of the same length (see for example [4]), for any two conjugate permutations $\varphi$ and $\psi$ we have $Z(\varphi)=Z(\psi)$. Obviously $Z(\varphi)=Z\left(\varphi^{-1}\right)$ for any permutation $\varphi$. So, $Z\left(\varphi_{i j}\right)=Z\left(\varphi_{j i}\right)$ for all spins.

Definition 2.4. Let $\Phi=\left\{\varphi_{1}, \varphi_{2}, \ldots, \varphi_{n}\right\}$ be a collection of permutations of the set $Q$. The set

$$
S p(\Phi)=\left[Z\left(\varphi_{1}\right), Z\left(\varphi_{2}\right), \ldots, Z\left(\varphi_{n}\right)\right]
$$

is called the spectrum of $\Phi$.
Two collections $\Phi=\left\{\varphi_{1}, \varphi_{2}, \ldots, \varphi_{n}\right\}$ and $\Sigma=\left\{\sigma_{1}, \sigma_{2}, \ldots, \sigma_{n}\right\}$ of permutations of $Q$ have the same spectrum if and only if there exists a permutation $\gamma$ of $Q$ such that $Z\left(\varphi_{i}\right)=Z\left(\sigma_{\gamma(i)}\right)$ for all $i=1,2, \ldots, n$.

The spectrum of all spins of a quasigroup $Q(\cdot)$, i.e., the set

$$
\left[Z\left(\varphi_{11}\right), Z\left(\varphi_{12}\right), \ldots, Z\left(\varphi_{n n}\right)\right]
$$

is called the spin-spectrum of $Q(\cdot)$ and is denoted by $\operatorname{Ssp}(Q, \cdot)$.
Theorem 2.5. Finite isotopic quasigroups have the same spin-spectrum.
Proof. Let $Q(\cdot)$ and $Q(\circ)$ be isotopic quasigroups. Then

$$
\gamma(x \circ y)=\alpha(x) \cdot \beta(y)
$$

for some permutations $\alpha, \beta, \gamma$ of $Q$.

In this case tracks of $Q(\cdot)$ and $Q(\circ)$ are connected by the formula (3). Spins of $Q(\cdot)$ and $Q(\circ)$ are pairwise conjugate. Namely

$$
\varphi_{\gamma(i) \gamma(j)}=\beta \psi_{i j} \beta^{-1} .
$$

Indeed,

$$
\begin{aligned}
\varphi_{\gamma(i) \gamma(j)} & =\varphi_{\gamma(i)} \varphi_{\gamma(j)}^{-1}=\left(\beta \psi_{i} \alpha^{-1}\right)\left(\beta \psi_{j} \alpha^{-1}\right)^{-1} \\
& =\left(\beta \psi_{i} \alpha^{-1}\right)\left(\alpha \psi_{j}^{-1} \beta^{-1}\right)=\beta \psi_{i} \psi_{j}^{-1} \beta^{-1}=\beta \psi_{i j} \beta^{-1} .
\end{aligned}
$$

Since spins $\varphi_{\gamma(i) \gamma(j)}$ and $\psi_{i j}$ are conjugate, we have $Z\left(\varphi_{\gamma(i) \gamma(j)}\right)=$ $Z\left(\psi_{i j}\right)$. This means that $Q(\cdot)$ and $Q(\circ)$ have the same spin-spectrum.

Corollary 2.6. If the isotopy of quasigroups $Q(\cdot)$ and $Q(\circ)$ has the form $(\alpha, \alpha, \gamma)$, then also sets of all $r$-tracks (l-tracks) of these quasigroups have the same spectrum.

Proof. Indeed, from (3) and (4), it follows that in this case $l$-tracks (respectively, $r$-tracks) of these quasigroups are pairwise conjugate.

## 3. Spin-basis of quasigroups

Definition 3.1. Let $\Phi$ be a collection of all nontrivial spins of a quasigroup $Q(\cdot)$. A minimal subset $B$ of $\Phi$ is called a basis of $\Phi$ if each spin from $\Phi$ can be written as a multiplication of spins (and their inverses) from $B$.

For example, the set

$$
B_{0}=\left\{\varphi_{12}, \varphi_{23}, \ldots, \varphi_{i(i+1)}, \ldots, \varphi_{(n-1) n}\right\}
$$

containing $(n-1)$ spins is a basis since each spin $\varphi_{p q}$, where $p<q$, can be written in the form

$$
\begin{aligned}
\varphi_{p q} & =\varphi_{p} \varphi_{q}^{-1}=\varphi_{p}\left(\varphi_{p+1}^{-1} \varphi_{p+1} \varphi_{p+2}^{-1} \varphi_{p+2} \ldots \varphi_{q-1}^{-1} \varphi_{q-1}\right) \varphi_{q}^{-1} \\
& =\left(\varphi_{p} \varphi_{p+1}^{-1}\right)\left(\varphi_{p+1} \varphi_{p+2}^{-1}\right) \ldots\left(\varphi_{q-1} \varphi_{q}^{-1}\right)=\varphi_{p(p+1)} \varphi_{(p+1)(p+2)} \cdots \varphi_{(q-1) q} .
\end{aligned}
$$

Also

$$
B_{i}=\left\{\varphi_{i 1}, \varphi_{i 2}, \ldots, \varphi_{i k}, \ldots, \varphi_{i n}\right\}, \quad i \neq k
$$

is a basis for every $i=1,2, \ldots, n$. Indeed, according to Proposition 2.3 (5), each spin $\varphi_{p q}$ can be written in the form

$$
\varphi_{p q}=\varphi_{i p}^{-1} \varphi_{i q} .
$$

Definition 3.2. Let $Q(\cdot)$ be a quasigroup of order $n$. The set

$$
\chi_{i}(Q, \cdot)=\left\{\varphi_{i 1}, \varphi_{i 2}, \ldots, \varphi_{i i}, \ldots, \varphi_{i n}\right\}=B_{i} \cup\left\{\varphi_{i i}\right\}
$$

is called the ith spin-basis of $Q(\cdot)$.
It coincides with the $i$ th row of the matrix $\left[\varphi_{i j}\right]$. In general, it is not closed under multiplication of spins, but in some cases it is a group. Since $\varphi_{k i} \varphi_{i j}=\varphi_{k j}$, by Proposition 2.3, for all $i, k=1,2, \ldots, n$ holds

$$
\varphi_{k i}\left(\chi_{i}(Q, \cdot)\right)=\chi_{k}(Q, \cdot)
$$

Proposition 3.3. If one of the spin-basis of a quasigroup $Q(\cdot)$ is a group, then each of its spin-basis is a group and

$$
\chi_{1}(Q, \cdot)=\chi_{2}(Q, \cdot)=\ldots=\chi_{n}(Q, \cdot) .
$$

Proof. Let $\chi_{i}(Q, \cdot)$ be a group. Then $\chi_{i}(Q, \cdot)$ together with $\varphi_{i k}$ contains also $\varphi_{i k}^{-1}=\varphi_{k i}$. This means that $\left\{\varphi_{1 i}, \varphi_{2 i}, \ldots, \varphi_{n i}\right\} \subseteq \chi_{i}(Q, \cdot)$. Therefore each spin $\varphi_{k j}$ belongs to $\chi_{i}(G, \cdot)$ because $\varphi_{k j}=\varphi_{k i} \varphi_{i j} \in \chi_{i}(G, \cdot)$ for all $j, k$. So, $\chi_{k}(Q, \cdot) \subseteq \chi_{i}(Q, \cdot)$ and $\varphi_{k i}\left(\chi_{i}(Q, \cdot)\right)=\chi_{k}(Q, \cdot)$ which completes the proof.

Proposition 3.4. Let quasigroups $Q(\cdot)$ and $Q(\circ)$ be isotopic. If one spinbasis of $Q(\cdot)$ is a group, then each spin-basis of $Q(\circ)$ is a group and for all $i=1, \ldots, n$ we have $\chi_{i}(Q, \cdot) \cong \chi_{i}(Q, \circ)$.

Proof. Let $\gamma(x \circ y)=\alpha(x) \cdot \beta(y)$. Then, as in the proof of Theorem 2.5,

$$
\varphi_{\gamma(i) \gamma(j)}=\beta \psi_{i j} \beta^{-1} .
$$

Whence

$$
\begin{equation*}
\psi_{i j}=\beta^{-1} \varphi_{\gamma(i) \gamma(j)} \beta . \tag{5}
\end{equation*}
$$

To prove that

$$
\chi_{i}(G, \circ)=\left\{\psi_{i 1}, \psi_{i 2}, \ldots, \psi_{i n}\right\}
$$

is a group observe that for all $\psi_{i p}, \psi_{i q} \in \chi_{i}(Q, \circ)$ we have

$$
\psi_{i p} \psi_{i q}=\beta^{-1} \varphi_{\gamma(i) \gamma(p)} \varphi_{\gamma(i) \gamma(q)} \beta=\beta^{-1} \varphi_{\gamma(i) k} \beta=\psi_{i t},
$$

where $\gamma(t)=k$, since, by Proposition 3.3, each spin-basis of $Q(\cdot)$ is a group. Moreover, for every $\psi_{i k} \in \chi_{i}(Q, \circ)$, by (5) and Proposition 2.3, we obtain

$$
\psi_{i k}^{-1}=\psi_{k i}=\beta^{-1} \varphi_{\gamma(k) \gamma(i)} \beta=\beta^{-1} \varphi_{\gamma(i) \gamma(k)}^{-1} \beta=\beta^{-1} \varphi_{\gamma(i) r} \beta=\psi_{i s},
$$

where $\gamma(s)=r$. This means that $\chi_{i}(Q, \circ)$ together with $\psi_{i k}$ also contains $\psi_{i k}^{-1}$. So, it is a group. Clearly $\chi_{i}(Q, \circ)=\chi_{k}(Q, \circ)$ for all $k=1, \ldots, n$.

In view of (5) the isomorphism $h: \chi_{\gamma(i)}(Q, \cdot) \rightarrow \chi_{i}(Q, \circ)=\chi_{\gamma(i)}(Q, \circ)$ has the form $h\left(\varphi_{\gamma(i) \gamma(j)}\right)=\beta^{-1} \varphi_{\gamma(i) \gamma(j)} \beta$.

Theorem 3.5. A finite quasigroup which is a group is isomorphic to its spin-basis.

Proof. Let $G(\cdot)$ be a group and $\chi_{1}(G, \cdot)=\left\{\varphi_{11}, \varphi_{12}, \ldots, \varphi_{1 n}\right\}$ its spin-basis. Then, according to the definition of spins, Proposition 1.1 and Corollary 1.2,

$$
\varphi_{1 i}(x)=\varphi_{1}\left(\lambda_{i}(x)\right)=\varphi_{1}\left(i \cdot x^{-1}\right)=\left(i \cdot x^{-1}\right)^{-1}=x \cdot i^{-1}=R_{i^{-1}}(x),
$$

which means that the spin-basis $\chi_{1}(G, \cdot)$ can be identified with the set of all right translations of $G(\cdot)$. So, $\chi_{1}(G, \cdot)$ and $G(\cdot)$ are isomorphic.

Proposition 3.3 completes the proof.
Theorem 3.6. A quasigroup for which the spin-basis is a group is isotopic to this group.

Proof. Let $Q(\circ)$ be a quasigroup. Since it is isotopic to some loop $Q(\cdot)$ with the identity 1, in view of Propositions 3.3 and 3.4 , it is sufficient to prove that $Q(\cdot)$ is isotopic to the group $\chi_{1}(Q, \cdot)=\left\{\varphi_{11}, \varphi_{12}, \varphi_{13}, \ldots, \varphi_{1 n}\right\}$.

For this we consider the mapping

$$
h: \chi_{1}(Q, \cdot) \longrightarrow Q(\cdot) \quad \text { such that } \quad h\left(\varphi_{1 i}\right)=i .
$$

It is one-to-one and onto. We prove that it is an isomorphism, i.e.,

$$
h\left(\varphi_{1 k} \varphi_{1 l}\right)=h\left(\varphi_{1 k}\right) \cdot h\left(\varphi_{1 l}\right)
$$

for all $\varphi_{1 k}, \varphi_{1 l}$ from $\chi_{i}(Q, \cdot)$.
As $\chi_{1}(Q, \cdot)$ is a group, the product of $\varphi_{1 k}$ and $\varphi_{1 l}$ also belongs to $\chi_{1}(Q, \cdot)$. Let

$$
\varphi_{1 k} \varphi_{1 l}=\varphi_{1 p} .
$$

By the definition of spins, the last equality is equivalent to

$$
\varphi_{1} \varphi_{k}^{-1} \varphi_{1} \varphi_{l}^{-1}=\varphi_{1} \varphi_{p}^{-1}
$$

i.e., to

$$
\varphi_{k}^{-1} \varphi_{1} \varphi_{l}^{-1}=\varphi_{p}^{-1}
$$

which can be written as

$$
\varphi_{p}=\varphi_{l} \varphi_{1}^{-1} \varphi_{k} .
$$

This means that

$$
\varphi_{p}(x)=\varphi_{l} \varphi_{1}^{-1} \varphi_{k}(x)
$$

holds for every $x \in Q$. Since $Q(\cdot)$ is a loop, the last identity is equivalent to

$$
x \cdot \varphi_{p}(x)=x \cdot \varphi_{l} \varphi_{1}^{-1} \varphi_{k}(x),
$$

whence, by (2), for $x=k$ we obtain

$$
p=k \cdot \varphi_{p}(k)=k \cdot \varphi_{l} \varphi_{1}^{-1} \varphi_{k}(k)=k \cdot \varphi_{l} \varphi_{1}^{-1}(1)=k \cdot \varphi_{l}(1)=k \cdot l
$$

because in any loop $\varphi_{k}(k)=1$ and $\varphi_{k}(1)=k$.
So, $h\left(\varphi_{1 k} \varphi_{1 l}\right)=p=k \cdot l=h\left(\varphi_{1 k}\right) \cdot h\left(\varphi_{1 l}\right)$, which completes the proof.
As a consequence of the above results we obtain
Theorem 3.7. A finite quasigroup is isotopic to a group if and only if its spin-basis is a group.

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# Semigroup, monoid and group models of groupoid identities 

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#### Abstract

In this note, we characterize those groupoid identities that have a (finite) non-trivial (semigroup, monoid, group) model.


## 1. Introduction

Definition 1.1. A groupoid consists of a non-empty set equipped with a binary operation, which we simply denote by juxtaposition. A groupoid $G$ is non-trivial if $|G|>1$, otherwise it is trivial. A semigroup is a groupoid $S$ that is associative $((x y) z=x(y z)$ for all $x, y, z \in S)$. A monoid is a semigroup $M$ possessing a neutral element $e \in M$ such that $e x=x e=x$ for all $x \in M$ (the letter $e$ will always denote the neutral element of a monoid). A group is a monoid $G$ such that for all $x \in G$ there exists an inverse $x^{-1}$ such that $x^{-1} x=x x^{-1}=e$. A quasigroup is a groupoid $Q$ such that for all $a, b \in Q$, there exist unique $x, y \in Q$ such that $a x=b$ and $y a=b$. A loop is a quasigroup possessing a neutral element.

A groupoid term is a product of universally quantified variables. A groupoid identity is an equation, the left-hand side and right-hand side of which are groupoid terms. By the words term and identity, we shall always mean groupoid term and groupoid identity, respectively. The letters $s$ and $t$ will always denote terms. We will say that an identity $s=t$ has a (finite) non-trivial model if there exists a (finite) non-trivial groupoid $G$ such that $s=t$ is valid in $G$. We will say that an identity $s=t$ has a (finite) nontrivial (semigroup, monoid, group, quasigroup, loop) model if $s=t$ has a

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(finite) non-trivial model that is a (semigroup, monoid, group, quasigroup, loop).

The question of whether or not an identity has a (finite) non-trivial model is known to be undecidable (not answerable by an algorithm) [3]. In this note, we show that the question of whether or not an identity has a (finite) non-trivial (semigroup, monoid, group) model is decidable.

## 2. Results

Lemma 2.1. An identity is valid in some non-trivial group if and only if it is valid in some non-trivial abelian group.

Proof. Suppose that the identity $s=t$ is valid in some non-trivial group $G$. Let $a$ be any non-neutral element of $G$. Then $s=t$ is valid in a non-trivial cyclic, and hence abelian, subgroup of $G$ containing $a$.

Given a term $t$ and a variable $x_{i}$, we denote by $o_{i}(t)$ the number of occurrences of $x_{i}$ in $t$. Given an identity $s=t$ and a variable $x_{i}$, we denote by $d_{i}$ the quantity $\left|o_{i}(s)-o_{i}(t)\right|$. Given an identity $s=t$ in the variables $x_{1}, x_{2}, \ldots, x_{n}$, we denote by $g$ the quantity $\operatorname{gcd}\left(d_{1}, d_{2}, \ldots, d_{n}\right)$.

Proposition 2.2. An identity $s=t$ in the variables $x_{1}, x_{2}, \ldots, x_{n}$ has $a$ non-trivial group model if and only if $g \neq 1$.

Proof. Suppose $g=1$. Suppose $s=t$ is valid in some non-trivial group $G$. By Lemma 2.1, $s=t$ is valid in some non-trivial abelian group $H$.

Now, in $H, s=t$ is equivalent to

$$
x_{1}^{d_{1}} x_{2}^{d_{2}} \cdots x_{n}^{d_{n}}=e .
$$

Let $m_{1}, m_{2}, \ldots, m_{n} \in \mathbb{Z}$ be such that $m_{1} d_{1}+m_{2} d_{2}+\cdots+m_{n} d_{n}=1$. Then, in $H$,

$$
\begin{aligned}
x & =x^{1}=x^{m_{1} d_{1}+m_{2} d_{2}+\cdots+m_{n} d_{n}}=x_{1}^{m_{1} d_{1}} x_{2}^{m_{2} d_{2}} \cdots x_{n}^{m_{n} d_{n}} \\
& =\left(x_{1}^{m_{1}}\right)^{d_{1}}\left(x_{2}^{m_{2}}\right)^{d_{2}} \cdots\left(x_{n}^{m_{n}}\right)^{d_{n}}=e,
\end{aligned}
$$

a contradiction.
Finally, suppose $g \neq 1$. Then $s=t$ is valid in the non-trivial group $\mathbb{Z}_{g}$.

As was mentioned before, the question of whether or not an identity has a finite non-trivial model is also known to be undecidable [3]. In fact, there exist identities with no non-trivial finite models but that do have infinite models, such as the identity $(((y y) y) x)(((y y)((y y) y)) z)=x[1]$.

Corollary 2.3. An identity has a non-trivial group model if and only if it has a finite non-trivial group model.

Proof. Suppose $s=t$ has a non-trivial group model. By Proposition 2.2, $g \neq 1$. Then $s=t$ is valid in the finite non-trivial group $\mathbb{Z}_{g}$.

Proposition 2.2 with "group" replaced by "loop" or "quasigroup" is false. Indeed, the identity $((x x) x) x=x(x x)$ is valid in the loop below (found with the model-generator Mace4 [2]).

| $\cdot$ | 0 | 1 | 2 | 3 | 4 | 5 | 6 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 0 | 1 | 2 | 3 | 4 | 5 | 6 |
| 1 | 1 | 2 | 3 | 5 | 0 | 6 | 4 |
| 2 | 2 | 4 | 5 | 1 | 6 | 3 | 0 |
| 3 | 3 | 0 | 6 | 4 | 2 | 1 | 5 |
| 4 | 4 | 3 | 0 | 6 | 5 | 2 | 1 |
| 5 | 5 | 6 | 1 | 0 | 3 | 4 | 2 |
| 6 | 6 | 5 | 4 | 2 | 1 | 0 | 3 |

It seems to be unknown if Corollary 2.3 with "group" replaced by "loop" or "quasigroup" is true.

Given the existence of a non-trivial idempotent $\left(x^{2}=x\right)$ monoid, Proposition 2.2 with "group" replaced by "monoid" is false. However, we now show that the question of whether or not an identity has a (finite) non-trivial monoid model is decidable.

Proposition 2.4. An identity $s=t$ in the variables $x_{1}, x_{2}, \ldots, x_{n}$ has a non-trivial monoid model if and only if every variable occurs on both sides or $g \neq 1$.

Proof. Suppose that there exists a variable $x$ that occurs $n>0$ times on one side of $s=t$ but not at all on the other side. Suppose $g=1$. Suppose that $s=t$ is valid in some non-trivial monoid $M$. Substituting $e$ for every variable in $s=t$ besides $x$ results in $x^{n}=e$. Therefore, every element of $M$ has an inverse and hence $M$ is a group. By Proposition 2.2, $M$ must be trivial, a contradiction.

Suppose that every variable in $s=t$ occurs on both sides. Then $s=t$ is valid in the non-trivial commutative idempotent monoid ( $G, \cdot \cdot$ ), where $G=\{0,1\}, 0 \cdot 0=0$ and $0 \cdot 1=1 \cdot 0=1 \cdot 1=1$.

Finally, suppose $g \neq 1$. Then $s=t$ is valid in the non-trivial group, and hence monoid, $\mathbb{Z}_{g}$.

Corollary 2.5. An identity has a non-trivial monoid model if and only if it has a non-trivial finite monoid model.

Proof. Suppose $s=t$ has a non-trivial monoid model. By Proposition 2.4, every variable that occurs in $s=t$ occurs on both sides or $g \neq 1$. If every variable that occurs in $s=t$ occurs on both sides, then $s=t$ is valid in the non-trivial commutative idempotent monoid above. If $g \neq 1$, then $s=t$ is valid in the finite non-trivial group, and hence monoid, $\mathbb{Z}_{g}$.

Proposition 2.4 with "monoid" replaced by "semigroup" is false. Indeed, $x y=z u$ is valid in a non-trivial zero semigroup and $x y=x(x y=y)$ is valid in a non-trivial left-zero (right-zero) semigroup. Nevertheless, we now show that the question of whether or not an identity has a (finite) non-trivial semigroup model is decidable.

Proposition 2.6. An identity $s=t$ in the variables $x_{1}, x_{2}, \ldots, x_{n}$ has a non-trivial semigroup model if and only if there are at least two variable occurrences on each side, one side is a variable which is also the left-most or right-most variable on the other side, or $g \neq 1$.

Proof. Suppose one side of $s=t$ is a variable $y$. Suppose $y$ is not the leftmost or right-most variable on the other side. Suppose $g=1$. Suppose $s=t$ is valid in some non-trivial semigroup $S$. Substituting $x$ for every variable in $s=t$ besides $y$ results in $x t(x, y) x=y$ for some (possibly empty) term $t(x, y)$ in the variables $x$ and $y$.

Now, in $S$,

$$
y t(y, x)(y t(y, z) y)=(y t(y, x) y) t(y, z) y .
$$

Therefore,

$$
y t(y, x) z=x t(y, z) y .
$$

Substituting $x$ for $y$ in the above results in

$$
x t(x, x) z=x t(x, z) x=z .
$$

Thus, $S$ is a monoid. By Proposition 2.4, $S$ must be trivial, a contradiction.
Suppose that there are at least two variable occurrences on each side of $s=t$. Then $s=t$ is valid in the non-trivial zero semigroup $(G, \cdot)$, where $G=\{0,1\}$ and $x \cdot y=0$.

Suppose one side of $s=t$ is a variable which is also the left-most (rightmost) variable on the other side. Then $s=t$ is valid in the non-trivial left-zero (right-zero) semigroup below.

$$
\begin{array}{c|cc}
\cdot & 0 & 1 \\
\hline 0 & 0 & 0 \\
1 & 1 & 1
\end{array}\left(\begin{array}{c|cc}
\cdot & 0 & 1 \\
\hline 0 & 0 & 1 \\
1 & 0 & 1
\end{array}\right)
$$

Suppose $g \neq 1$. Then $s=t$ is valid in the non-trivial group, and hence semigroup, $\mathbb{Z}_{g}$.

Corollary 2.7. An identity has a non-trivial semigroup model if and only if it has a finite non-trivial semigroup model.

Proof. Suppose $s=t$ has a non-trivial semigroup model. By Proposition 2.6 , there are at least two variable occurrences on each side of $s=t$, one side of $s=t$ is a variable which is also the left-most or right-most variable on the other side, or $g \neq 1$. If there are at least two variable occurrences on each side of $s=t$, then $s=t$ is valid in the finite non-trivial zero semigroup above. If one side of $s=t$ is a variable which is also the left-most (rightmost) variable on the other side, then $s=t$ is valid in the finite non-trivial left-zero (right-zero) semigroup above. If $g \neq 1$, then $s=t$ is valid in the finite non-trivial group, and hence semigroup, $\mathbb{Z}_{g}$.

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# Direct product of quasigroups and generalized diagonal subquasigroup 

Tuval Foguel


#### Abstract

In this paper we look at when the direct product $\mathcal{P} \times \mathcal{Q}$ of two quasigroups contains a subquasigroup isomorphic to $\mathcal{P}$.


## 1. Introduction

The direct product $\mathcal{P} \times \mathcal{Q}$ of two groups (loops) clearly contains at least one subgroup (subloop) isomorphic to $\mathcal{P}$, namely $\mathcal{P} \times\{\mathbf{1}\}$. This is not the case for a direct product of two quasigroups. Bruck in [4] gives examples of finite nontrivial quasigroups $\mathcal{P}$ and $\mathcal{Q}$ whose direct product has no proper subquasigroups.

In this paper we will look at what we can say about the quasigroups $\mathcal{P}$ and $\mathcal{Q}$ if their direct product contains a subquasigroup isomorphic to $\mathcal{P}$.

## 2. Preliminaries

In this section, we review a few necessary notions from quasigroup theory and establish some notation conventions.

A magma $(\mathcal{Q}, \cdot)$ consists of a set $\mathcal{Q}$ together with a binary operation on $\mathcal{Q}$. For $x \in \mathcal{Q}$, define the left (resp., right) translation by $x$ by $L(x) y=x y$ (resp., $R(x) y=y x)$ for all $y \in \mathcal{Q}$. A magma with all left and right translations bijective is called a quasigroup. A quasigroup $\mathcal{Q}$ is an idempotent quasigrop if for all $x \in \mathcal{Q}, x x=x$. A quasigroup $\mathcal{L}$ with a two-sided identity element $\mathbf{1}$ such that for any $x \in \mathcal{L}, x \mathbf{1}=\mathbf{1} x=x$ is called a loop. A loop $\mathcal{L}$ is power-associative, if for any $x \in \mathcal{L}$, the subloop generated by $x$ is a
group. For basic facts about loops and quasigroups, we refer the reader to [2], [3] and [7].

Notation 2.1. Given the direct product $\mathcal{P} \times \mathcal{Q}$ of two quasigroups, we will denote the $i^{\text {th }}$ projection homomorpishm by $\pi_{i}$.

Notation 2.2. Given two quasigroups $\mathcal{K}$ and $\mathcal{Q}$, we will denote that $\mathcal{K}$ is a subquasigroup of $\mathcal{Q}$ by $\mathcal{K} \leq \mathcal{Q}$, and that $\mathcal{K}$ is a subquasigroup of $\mathcal{Q}$ but not equal to $\mathcal{Q}$ by $\mathcal{K} \leq \mathcal{Q}$.

## 3. Generalized diagonal subquasigroup

Lemma 3.1. If $\hat{\mathcal{Q}}$ is a homomorphic image of a quasigroup $\mathcal{P}$ and $\hat{\mathcal{Q}} \subseteq \mathcal{Q}$ a quasigroup, then $\hat{\mathcal{Q}}$ is a quasigroup.

Proof. See [3].
Lemma 3.2. $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups if and only if there exists a homomorphism $f: \mathcal{P} \rightarrow \mathcal{Q}$.

Proof. Assume $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$. Then $\pi_{2}$ is a homomorphism from $\mathcal{K} \rightarrow \mathcal{Q}$ and since $\mathcal{P} \cong \mathcal{K}$ there exists a homomorphism $f: \mathcal{P} \rightarrow \mathcal{Q}$. Conversely, if there exists a homomorphism $f: \mathcal{P} \rightarrow \mathcal{Q}$, then $\mathcal{P} \cong\{(p, f(p)) \mid p \in \mathcal{P}\} \leq$ $\mathcal{P} \times \mathcal{Q}$.

Corollary 3.3. $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups if and only if $\mathcal{Q}$ contains a subquasigroup that is a homomorphic image of $\mathcal{P}$.

Proof. See Lemma 3.1 and Lemma 3.2.
Corollary 3.4. $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups with $\mathcal{Q}$ containing no subquasigroups except for itself if and only if $\mathcal{Q}$ is a homomorphic image of $\mathcal{P}$.

Definition 3.5. Given $\mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups, we will call a subquasigroup $\mathcal{K}$ a generalized diagonal subquasigroup (gd-subquasigroup) if $\mathcal{K}=\{(p, f(p)) \mid p \in \mathcal{P}\} \leq \mathcal{P} \times \mathcal{Q}$ where $f$ is a homomorphism from $\mathcal{P}$ to $\mathcal{Q}$.

Example 3.6. If $\mathcal{P}$ and $\mathcal{Q}$ are loops, then $\mathcal{K}=\mathcal{P} \times\{\mathbf{1}\} \leq \mathcal{P} \times \mathcal{Q}$ is a gd-subquasigroup.

Example 3.7. The diagonal-subquasigroup $\{(p, p) \mid p \in \mathcal{P}\}$ is a gd-subquasigroup of $\mathcal{P} \times \mathcal{P}$.

Theorem 3.8. $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups if and only if $\mathcal{P} \times \mathcal{Q}$ contains a gd-subquasigroup.

Proof. By the definition of a gd-subquasigroup, it is isomorphic to $\mathcal{P}$.
If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$, then by Lemma 3.2 there is a homomorphism $f: \mathcal{P} \rightarrow \mathcal{Q}$. Thus $\{(p, f(p)) \mid p \in \mathcal{P}\} \leq \mathcal{P} \times \mathcal{Q}$ is a gd-subquasigroup of $\mathcal{P} \times \mathcal{Q}$.

Definition 3.9. A quasigroup is said to have a covering by subquasibgroups if it is the set-theoretic union of proper subquasigroups, and, if the set of subquasigroups is finite, we say the covering is finite. Such coverings have been widely studied in groups, and recently, analogous coverings for rings, semigroups, and loops have been discussed in [1], [6], and [5], respectively. A covering is disjoint if any two distinct subquasigroups in the covering are disjoint.

Lemma 3.10. If $\mathcal{P}$ is a quasigroup and $\mathcal{Q}$ is an idempotent quasigroup, then $\mathcal{P} \times \mathcal{Q}$ has a disjoint covering $\mathcal{P} \times \mathcal{Q}=\bigcup_{i \in \mathcal{Q}}(\mathcal{P} \times\{i\})$ where $\mathcal{P} \times\{i\} \cong \mathcal{P}$ for all $i \in \mathcal{Q}$

Proof. $\mathcal{P} \times\{i\} \cong \mathcal{P}$ since $i$ is an idempotent for all $i \in \mathcal{Q}$. If $i, j \in \mathcal{Q}$ and $i \neq j$, then $\mathcal{P} \times\{i\} \bigcap \mathcal{P} \times\{j\}=\emptyset$ and if $h \in \mathcal{P} \times \mathcal{Q}$, then $h=(p, i)$ where $p \in \mathcal{P}$ and $i \in\{i\} \leq \mathcal{Q}$.

Definition 3.11. A quasigroup is homogeneous if its automorphism group is transitive. A quasigroup is doubly homogeneous if its automorphism group is doubly transitive. A two-quasigroup is a nontrivial two generated doubly homogeneous quasigroup.

Remark 3.12. If $\mathcal{Q}$ is a two-quasigroup, then it is generated as a quasigroup by any two distinct elements, and by [8] $\mathcal{Q}$ is an idempotent quasigroup.

Example 3.13. Given $\mathcal{Q}=G F\left(p^{n}\right)$ (the Galois field of $p^{n}$ elements), and $\alpha$ a primitive element in $G F\left(p^{n}\right)$. Then $(\mathcal{Q}, \odot)$ is a two-qusigroup under the binary operation

$$
a \odot b=\alpha a+(\mathbf{1}-\alpha) b
$$

for all $a, b \in \mathcal{Q}$.
Lemma 3.14. If $\mathcal{P}$ is a quasigroup with no subquasigroups except for itself, and $\mathcal{Q}$ is a two-quasigroup, then every proper subquasigroup of $\mathcal{P} \times \mathcal{Q}$ is of the form $\mathcal{P} \times\{i\} \cong \mathcal{P}$ where $i \in \mathcal{Q}$.

Proof. Assume $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$. Since $\pi_{1}(\mathcal{K}) \leq \mathcal{P}$ and $\mathcal{P}$ is a quasigroup with no proper subquasigroups, $\pi_{1}(\mathcal{K})=\mathcal{P}$.

Let $k_{1}=\left(p_{1}, i\right), k_{2}=\left(p_{2}, j\right) \in \mathcal{K}$. If $i \neq j$, then since $\mathcal{Q}$ is a twoquasigroup $\pi_{2}(\mathcal{K})=\mathcal{Q}$. Therefore given any $(p, t) \in \mathcal{P} \times \mathcal{Q}$ there exist $k=(\hat{p}, t) \in \mathcal{K}$, but some "power" of $\hat{p}$ is equal to $p$, and thus $\mathcal{K}=\mathcal{P} \times \mathcal{Q}$. So if $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$, then $\mathcal{K}=\mathcal{P} \times \mathcal{Q}$ or $\mathcal{K}=\mathcal{P} \times\{i\}$.

## 4. The non gd-subquasigroup

Lemma 3.1. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups, then $\pi_{1}(\mathcal{K}) \leq \mathcal{P}$ and $\pi_{1}(\mathcal{K})$ is a homomrphic image of $\mathcal{P}$.

Proof. $\pi_{1}(\mathcal{K}) \leq \mathcal{P}$ by definition. Since $\pi_{1}(\mathcal{K})$ is a homomrphic image of $\mathcal{K}$ it is a homomrphic image of $\mathcal{P} \cong \mathcal{K}$.

Corollary 3.2. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a simple quasigroup and $\mathcal{Q}$ is a quasigroup, then $\pi_{1}(\mathcal{K}) \cong \mathcal{P}$ or $\{\mathbf{1}\}$.

Corollary 3.3. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a quasigroup with no subquasigroups except for itself, and $\mathcal{Q}$ is a quasigroup, then $\pi_{1}(\mathcal{K}) \cong \mathcal{P}$.

Example 3.4. In $\mathbb{Z} \times \mathcal{L}$, where $\mathbb{Z}$ denotes the integers and $\mathcal{L}$ is any loop, $\mathcal{K}=2 \mathbb{Z} \times\{\mathbf{1}\} \cong \mathbb{Z} \cong \pi_{1}(\mathcal{K})$, but note that $\pi_{1}(\mathcal{K}) \neq \mathbb{Z}$.

Remark 3.5. Given $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups and $\mathcal{P}$ is finite, $\pi_{1}(\mathcal{K}) \cong \mathcal{P}$ if and only if $\pi_{1}(\mathcal{K})=\mathcal{P}$.

Definition 3.6. Given nonempty subsets $A$ and $B$ of a quasigroup $\mathcal{P}$, we will denote by $A B=\{a b \mid a \in A, b \in B\}$.

The following deffinition is due to Bruck (see [4]).
Definition 3.7. Let $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups. For $p \in \mathcal{P}$ denote by $\mathcal{Q}_{p}=\{q \in \mathcal{Q} \mid(p, q) \in \mathcal{K}\} \subseteq \mathcal{Q}$.

Lemma 3.8. If $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups, then $\mathcal{Q}_{p} \mathcal{Q}_{\grave{p}}=$ $\mathcal{Q}_{p \grave{p}}$ for $p, \grave{p} \in \pi_{1}(\mathcal{K})$.

Proof. See [4] Lemma 15. Note that finiteness is not used in this part of the proof.

Remark 3.9. If $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups and $p \in$ $\mathcal{P}-\pi_{1}(\mathcal{K})$, then $\mathcal{Q}_{p}=\emptyset$.

Lemma 3.10. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a loop and $\mathcal{Q}$ is a quasigroup, then $\mathcal{Q}_{\mathbf{1}}$ is isomorphic to a normal subloop of $\mathcal{P}$

Proof. $\mathcal{Q}_{1}$ is isomorphic to the kernel of $\pi_{1}$.
Lemma 3.11. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a finite power associative loop and $\mathcal{Q}$ is a quasigroup, then $\underbrace{\mathcal{Q}_{p} \cdots \mathcal{Q}_{p}}_{|p| \text { times }}=\mathcal{Q}_{\mathbf{1}}$ for any $p \in \pi_{1}(\mathcal{K})$.

Proof. By Lemma $3.8 \underbrace{\mathcal{Q}_{p} \cdots \mathcal{Q}_{p}}_{|p|-\text { times }}=\mathcal{Q}_{p|p|}=\mathcal{Q}_{\mathbf{1}}$.
Definition 3.12. For a finite power associative loop $\mathcal{P}, \exp (\mathcal{P})=n$ is the smallest positive integer such that given $p \in \mathcal{P}$ the identity $p^{n}=\mathbf{1}$ holds.

Corollary 3.13. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a finite power associative loop and $\mathcal{Q}$ is a quasigroup, then for any $q \in \pi_{2}(\mathcal{K}), q^{\exp (\mathcal{P})} \in \mathcal{Q}_{\mathbf{1}}$.
Proof. $q \in \mathcal{Q}_{p}$ for some $p \in \pi_{1}(\mathcal{K})$, and thus $q^{\exp (\mathcal{P})} \in \mathcal{Q}_{p} \exp (\mathcal{P})=\mathcal{Q}_{\mathbf{1}}$.
Remark 3.14. If $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups and $\mathcal{K}$ is finite, then $|\mathcal{K}|=\sum_{p \in \mathcal{P}}\left|\mathcal{Q}_{p}\right|=\sum_{p \in \pi_{1}(\mathcal{K})}\left|\mathcal{Q}_{p}\right|$.
Lemma 3.15. If $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups and $\mathcal{K}$ is finite, then $\left|\mathcal{Q}_{p}\right|=\left|\mathcal{Q}_{\grave{p}}\right|$ for $p, \grave{p} \in \pi_{1}(\mathcal{K})$ and $\mathcal{Q}_{p}$ and $\mathcal{Q}_{\grave{p}}$ are either disjoint or identical.

Proof. By Remark 3.14 we see that $\left|\mathcal{Q}_{p}\right|$ is finite for each $p$, and thus by [4] Lemma $15,\left|\mathcal{Q}_{p}\right|=\left|\mathcal{Q}_{\grave{p}}\right|$ for $p, \grave{p} \in \pi_{1}(\mathcal{K})$ where $\mathcal{Q}_{p}$ and $\mathcal{Q}_{\grave{p}}$ are either disjoint or identical.

Remark 3.16. If $\mathcal{K} \cong \mathcal{P} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a finite quasigroup and $\mathcal{Q}$ is quasigroup, then $|\mathcal{P}|=\sum_{p \in \mathcal{P}}\left|\mathcal{Q}_{p}\right|=\sum_{p \in \pi_{1}(\mathcal{K})}\left|\mathcal{Q}_{p}\right|$.
Lemma 3.17. If $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups and $\mathcal{K}$ is finite, then $\left|\pi_{1}(\mathcal{K})\right|\left|\mathcal{Q}_{p}\right|=|\mathcal{K}|$ for any $p \in \pi_{1}(\mathcal{K})$.
Proof. By Remarks 3.16 and Lemma 3.15 we get that

$$
|\mathcal{K}|=\sum_{p \in \pi_{1}(\mathcal{K})}\left|\mathcal{Q}_{p}\right|=\left|\pi_{1}(\mathcal{K})\right|\left|\mathcal{Q}_{p}\right| .
$$

Corollary 3.18. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a finit quasigroup and $\mathcal{Q}$ is quasigroup, then $\left|\pi_{1}(\mathcal{K})\right|\left|\mathcal{Q}_{p}\right|=|\mathcal{P}|$ for any $p \in \pi_{1}(\mathcal{K})$.

Lemma 3.19. If $\mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups, $\pi_{1}(\mathcal{K}) \cong \mathcal{K}$ and $\mathcal{K}$ is finite, then $p \mapsto \mathcal{Q}_{p}$ for all $p \in \pi_{1}(\mathcal{K})$ is a homomorphism from $\pi_{1}(\mathcal{K})$ to $\mathcal{Q}$.
Proof. By Lemma 3.17 we see that $|\mathcal{K}|=\left|\pi_{1}(\mathcal{K})\right|\left|\mathcal{Q}_{p}\right|=|\mathcal{K}|\left|\mathcal{Q}_{p}\right|$, and thus we get that $\left|\mathcal{Q}_{p}\right|=1$ for all $p \in \pi_{1}(\mathcal{K})$. Therefore by Lemma $3.8 p \mapsto \mathcal{Q}_{p}$ is a homomorphism for all $p \in \pi_{1}(\mathcal{K})$.

Corollary 3.20. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ is a finite quasigroup, $\mathcal{Q}$ is quasigroup, and $\pi_{1}(\mathcal{K}) \cong \mathcal{P}$, then $p \mapsto \mathcal{Q}_{p}$ for all $p \in \mathcal{P}$ is a homomorphism from $\mathcal{P}$ to $\mathcal{Q}$ and $\mathcal{K}$ is a gd-subquasigroup.
Proof. Note that $\mathcal{K}=\left\{\left(p, \mathcal{Q}_{p}\right) \mid p \in \mathcal{P}\right\}$.
Theorem 3.21. If $\mathcal{P} \cong \mathcal{K} \leq \mathcal{P} \times \mathcal{Q}$ where $\mathcal{P}$ and $\mathcal{Q}$ are quasigroups with $\mathcal{P}$ finite, then $\pi_{1}(\mathcal{K}) \leq \mathcal{P}, \pi_{1}(\mathcal{K})$ is a homomrphic image of $\mathcal{P},\left|\mathcal{Q}_{p}\right|\left|\pi_{1}(\mathcal{K})\right|=$ $|\mathcal{P}|$ for any $p \in \pi_{1}(\mathcal{K})$, and if $\mathcal{P}=\pi_{1}(\mathcal{K})$, then $\mathcal{K}$ is a gd-subquasigroup.
Proof. This follows from Lemmas 3.1, Corollary 3.18 and Corollay 3.20.
Example 3.22. Let $\mathcal{Q}$ be a finite quasigroup, $\mathcal{P}=\mathcal{Q} \times \mathcal{Q}$, and $\mathcal{K}=$ $\{(q, q, \hat{q}) \mid q, \hat{q} \in \mathcal{Q}\} \subseteq \mathcal{P} \times \mathcal{Q}$. Then $\mathcal{K} \cong \mathcal{P}$ but $\pi_{1}(\mathcal{K}) \not \not \mathcal{P}$ and $\pi_{2}(\mathcal{K}) \not \not \mathcal{P}$.
Example 3.23. Let $\mathcal{P}=\mathcal{L} \times \mathbb{Z}_{n}=\mathcal{Q}$ where $\mathcal{L}$ is a loop, $\mathbb{Z}_{n}$ denotes the integers mod $n$, and let $\mathcal{K}=\left\{(l, 0, l, i) \mid l \in \mathcal{L}\right.$ and $\left.i \in \mathbb{Z}_{n}\right\}$. Then $\mathcal{Q}_{(l, 0)}=\left\{(l, i) \mid i \in \mathbb{Z}_{n}\right\}$ is not a quasigroup if $l \neq \mathbf{1}$ and $\left|\mathcal{Q}_{(l, 0)}\right|=n$.

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# Algebraic properties of some varieties of central loops 

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#### Abstract

Isotopes of C-loops with a unique non-identity squares are studied. It is proved that such loops are C-loops and A-loops. The relationship between C-loops and Steiner loops is further studied. Central loops with the weak and cross inverse properties are also investigated.


## 1. Introduction

C-loops are one of the least studied loops. Few publications that have considered C-loops include Fenyves [14], [15], Beg [7], [8], Phillips et. al. [24], [26], [21], [20], Chein [10] and Solarin et. al. [2], [30], [28], [27]. The difficulty in studying them is as a result of the nature of their identities when compared with other Bol-Moufang identities (the element occurring twice on both sides has no other element separating it from itself). Latest publications on the study of C-loops which has attracted fresh interest on the structure include [24], [26], and [21].
$L C$-loops, $R C$-loops and $C$-loops are loops that satisfies the identities

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

respectively. Fenyves' work in [15] was completed in [24]. Fenyves proved that LC-loops and RC-loops are defined by three equivalent identities. In [24] and [25], it was shown that LC-loops and RC-loops are defined by four equivalent identities. Solarin [28] named the fourth identities the left middle ( $L M$ ) and right middle ( $R M$ ) identities and loops that obey them are called

[^4]LM-loops and RM-loops, respectively. These terminologies were also used in [29]. Their basic properties are found in [26], [15] and [13].

The right and left translation on a loop ( $L, \cdot$ ) are bijections $R_{x}: L \rightarrow L$ and $L_{x}: L \rightarrow L$ defined as $y R_{x}=y x$.

Definition 1.1. Let $(L, \cdot)$ be a loop. The left nucleus of $L$ is the set

$$
N_{\lambda}(L, \cdot)=\{a \in L: a x \cdot y=a \cdot x y \forall x, y \in L\} .
$$

The right nucleus of $L$ is the set

$$
N_{\rho}(L, \cdot)=\{a \in L: y \cdot x a=y x \cdot a \forall x, y \in L\}
$$

The middle nucleus of $L$ is the set

$$
N_{\mu}(L, \cdot)=\{a \in L: y a \cdot x=y \cdot a x \forall x, y \in L\} .
$$

The nucleus of $L$ is the set

$$
N(L, \cdot)=N_{\lambda}(L, \cdot) \cap N_{\rho}(L, \cdot) \cap N_{\mu}(L, \cdot) .
$$

The centrum of $L$ is the set

$$
C(L, \cdot)=\{a \in L: a x=x a \forall x \in L\} .
$$

The center of $L$ is the set

$$
Z(L, \cdot)=N(L, \cdot) \cap C(L, \cdot) .
$$

$L$ is said to be a centrum square loop if $x^{2} \in C(L, \cdot)$ for all $x \in L$. $L$ is said to be a central square loop if $x^{2} \in Z(L, \cdot)$ for all $x \in L . L$ is said to be left alternative if for all $x, y \in L, x \cdot x y=x^{2} y$ and is said to right alternative if for all $x, y \in L, y x \cdot x=y x^{2}$. Thus, $L$ is said to be alternative if it is both left and right alternative. The triple ( $U, V, W$ ) such that $U, V, W \in \operatorname{SYM}(L, \cdot)$ is called an autotopism of $L$ if and only if

$$
x U \cdot y V=(x \cdot y) W \quad \forall x, y \in L .
$$

$\operatorname{SYM}(L, \cdot)$ is called the permutation group of the loop $(L, \cdot)$. The group of autotopisms of $L$ is denoted by $A U T(L, \cdot)$. Let $(L, \cdot)$ and $(G, \circ)$ be two distinct loops. The triple $(U, V, W):(L, \cdot) \rightarrow(G, \circ)$ such that $U, V, W$ : $L \rightarrow G$ are bijections is called a loop isotopism if and only if

$$
x U \circ y V=(x \cdot y) W \quad \forall x, y \in L
$$

We investigate central loops with the unique non-identity commutators, associators and squares. The relationship between C-loops and Steiner loops is studied. Central loops with the weak and cross inverse properties are also investigated.

For definition of concepts in theory of loops readers may consult [9], [29] and [23].

## 2. Preliminaries

Definition 2.1. (cf. [16]) Let $a, b$ and $c$ be three elements of a loop $L$. The loop commutator of $a$ and $b$ is the unique element $(a, b)$ of $L$ such that $a b=(b a)(a, b)$. The loop associator of $a, b$ and $c$ is the unique element $(a, b, c)$ of $L$ such that $(a b) c=\{a(b c)\}(a, b, c)$.

If $X, Y$, and $Z$ are subsets of a loop $L$, we denote by $(X, Y)$ and $(X, Y, Z)$, respectively, the set of all commutators of the form $(x, y)$ and all the associators of the form $(x, y, z)$, where $x \in X, y \in Y, z \in Z$.

Definition 2.2. (cf. [16]) A unique non-identity commutator is an element $s \neq e$ ( $e$ is the identity element) in a loop $L$ with the property that every commutator in $L$ is $e$ or $s$. A unique non-identity commutator associator is an element $s \neq e$ in a loop $L$ with the property that every commutator in $L$ is $e$ or $s$ and every associator is $e$ or $s$. A unique non-identity square or non-trivial square is an element $s \neq e$ in a loop $L$ with the property that every square in $L$ is $e$ or $s$.

Definition 2.3. A loop ( $L, \cdot$ ) is called a weak inverse property loop (W.I.P.L.) if and only if it satisfies the weak inverse property (W.I.P.): $y(x y)^{\rho}=x^{\rho}$ for all $x, y \in L$. $L$ is called a cross inverse property loop (C.I.P.L.) if and only if it satisfies the cross inverse property (C.I.P.): $x y \cdot x^{\rho}=y .(L, \cdot)$ is a left (right) inverse property loop (L.I.P.L.) (resp. (R.I.P.L.)) if and only if it has the left (resp. right) inverse property (L.I.P) (resp. (R.I.P)): $x^{\lambda}(x y)=y$ $\left(\operatorname{resp}(y x) x^{\rho}=y\right.$. It is an inverse property loop (I.P.L.) if and only if it has the inverse property (I.P.) i.e., it has L.I.P. and R.I.P. property.

Most of our results and proofs, are written in dual form relative to RCloops and LC-loops. That is, a statement like 'LC(RC)-loop... A(B)' where 'A' and 'B' are some equations or expressions means that ' A ' is for LC-loops and ' B ' is for RC -loops.

## 3. Inner mappings

Lemma 3.1. Let $L$ be a C-loop. Then for each $(A, B, C) \in A U T(L)$, there exists a unique pair $\left(S_{1}, T_{1}, R_{1}\right),\left(S_{2}, T_{2}, R_{2}\right) \in \operatorname{AUT}(L, \cdot)$ such that $L_{x}^{2}=S_{2}^{-1} S_{1}, R_{x}^{2}=T_{1}^{-1} T_{2}, R_{x}^{-2} L_{x}^{2}=R_{2}^{-1} R_{1}, R_{1}^{-1} R_{2} T_{2}^{-1} T_{1} S_{2}^{-1} S_{1}=I$ for all $x \in L$.

Proof. If $L$ is a C-loop, then $\left(L_{x}^{2}, I, L_{x}^{2}\right),\left(I, R_{x}^{2}, R_{x}^{2}\right) \in A U T(L)$ for all $x \in L$. So, there exist $\left(S_{1}, T_{1}, R_{1}\right),\left(S_{2}, T_{2}, R_{2}\right) \in A U T(L)$ such that

$$
\begin{aligned}
& \left(S_{1}, T_{1}, R_{1}\right)=(A, B, C)\left(L_{x}^{2}, I, L_{x}^{2}\right) \in \operatorname{AUT}(L) \\
& \left(S_{2}, T_{2}, R_{2}\right)=(A, B, C)\left(I, R_{x}^{2}, R_{x}^{2}\right) \in \operatorname{AUT}(L) .
\end{aligned}
$$

Hence, the conditions hold although the identities do not depend on $(A, B, C)$, but the uniqueness does.

Theorem 3.1. Let $L$ be a C-loop and let there exist a unique pair of autotopisms $\left(S_{1}, T_{1}, R_{1}\right),\left(S_{2}, T_{2}, R_{2}\right)$ such that the conditions $L_{x}^{2}=S_{2}^{-1} S_{1}$, $R_{x}^{2}=T_{1}^{-1} T_{2}$ and $R_{x}^{-2} L_{x}^{2}=R_{2}^{-1} R_{1}$ hold for each $x \in L$. If $\alpha_{1}=S_{1}^{-1}$, $\alpha_{2}=S_{2}^{-1}, \beta_{1}=T_{1}^{-1}, \beta_{2}=T_{2}^{-1}, \gamma_{1}=R_{1}^{-1}$ and $\gamma_{2}=R_{2}^{-1}$, then

$$
\left(x^{2} y\right) \alpha_{1}=y \alpha_{2}, \quad\left(y x^{2}\right) \beta_{2}=y \beta_{1}, \quad\left(x^{2} y x^{-2}\right) \gamma_{1}=y \gamma_{2} \forall x, y \in L .
$$

Proof. From Lemma 3.1 we have $L_{x}^{2}=S_{2}^{-1} S_{1}, R_{x}^{2}=T_{1}^{-1} T_{2}, R_{x}^{-2} L_{x}^{2}=$ $R_{2}^{-1} R_{1}$. Keeping in mind that a C-loop is power associative and nuclear square, we have the following:

1. $L_{x}^{2}=S_{2}^{-1} S_{1} \longleftrightarrow y L_{x}^{2}=y S_{2}^{-1} S_{1}$ for all $y \in L \longleftrightarrow y L_{x^{2}}=y S_{2}^{-1} S_{1} \longleftrightarrow$ $x^{2} y=y S_{2}^{-1} S_{1} \longleftrightarrow\left(x^{2} y\right) S_{1}^{-1}=y S_{2}^{-1} \longleftrightarrow x^{2} y \alpha_{1}=y \alpha_{2}$.
2. $R_{x}^{2}=T_{1}^{-1} T_{2} \longleftrightarrow y R_{x}^{2}=y T_{1}^{-1} T_{2}$ for all $y \in L \longleftrightarrow y x^{2}=y T_{1}^{-1} T_{2}$
$\longleftrightarrow y x^{2} T_{2}^{-1}=y T_{1}^{-1} \longleftrightarrow y x^{2} \beta=y \beta_{1}$.
3. $R_{x}^{-2} L_{x}^{2}=R_{2}^{-1} R_{1} \longleftrightarrow y R_{x}^{-2} L_{x}^{2}=y R_{2}^{-1} R_{1}$ for all $y \in L \longleftrightarrow x^{2} y x^{-2}=$ $y R_{2}^{-1} R_{1} \longleftrightarrow\left(x^{2} y x^{-2}\right) R_{1}^{-1}=y R_{2}^{-1} \longleftrightarrow\left(x^{2} y x^{-2}\right) \gamma_{1}=y \gamma_{2}$.

Corollary 3.1. Let L be a C-loop. An autotopism of $L$ can be constructed if there exists at least one $x \in L$ such that $x^{2} \neq e$. In this case also the inverse can be constructed.

Proof. We need Lemma 3.1 and Theorem 3.1. If $x^{2}=e$, then the autotopism is trivial. Since $L$ is a C-loop, using Lemma 3.1 and Theorem 3.1, it will be noticed that $\left(\alpha_{1} S_{2}, \beta_{1} T_{2}, \gamma_{1} R_{2}\right) \in \operatorname{AUT}(L)$ and $\left(\alpha_{2} S_{1}, \beta_{2} T_{1}, \gamma_{2} R_{1}\right)=$ $\left(\alpha_{1} S_{2}, \beta_{1} T_{2}, \gamma_{1} R_{2}\right)^{-1}$. Hence the proof.

Lemma 3.2. For a C-loop $L$ the mapping $\gamma_{2} R_{1}: L \rightarrow L$ used in the autotopism $\left(\alpha_{2} S_{1}, \beta_{2} T_{1}, \gamma_{2} R_{1}\right) \in A U T(L)$ and defined by the identity $y \gamma_{2} R_{1}=x^{2} y x^{-2}$ for all $x \in L$ is:

1. an automorphism,
2. a semi-automorphism,
3. a middle inner mapping,
4. a pseudo-automorphism with companion $x^{2}$.

Proof. 1. The map $\gamma_{2} R_{1}$ is a bijection by the construction of the autotopism $\left(\alpha_{2} S_{1}, \beta_{2} T_{1}, \gamma_{2} R_{1}\right) \in \operatorname{AUT}(L)$. So we need only to show that it is an homomorphism. Let $y_{1}, y_{2} \in L$, then: $\left(y_{1} y_{2}\right) \gamma_{2} R_{1}=\left(x^{2} y_{1} x^{-2}\right)\left(x^{2} y_{2} x^{-2}\right)=$ $y_{1} \gamma_{2} R_{1} \cdot y_{2} \gamma_{2} R_{1}$. Whence, $\gamma_{2} R_{1}$ is an automorphism.
2. We have $e \gamma_{1}=e \gamma_{2}$, hence $e \gamma_{2} R_{1}=e$. Thus $(z y \cdot z) \gamma_{2} R_{1}=x^{2}(z y \cdot z) x^{-2}=$ $x^{2}\left((z y \cdot z) x^{-2}\right)=\left(x^{2} z x^{-2}\right)\left(x^{2} y x^{-2}\right) \cdot z \gamma_{2} R_{1}=\left(z \gamma_{2} R_{1} \cdot y \gamma_{2} R_{1}\right) \cdot z \gamma_{2} R_{1}$. So, $\gamma_{2} R_{1}$ is a semi-automorphism.
3. Since $e \gamma_{2} R_{1}=e$, we hawe $y \gamma_{2} R_{1}=y R_{x^{-2}} L_{\left(x^{-2}\right)^{-1}}=y T\left(x^{-2}\right)$ for all $y \in L$, which implies $\gamma_{2} R_{1}=T\left(x^{-2}\right) \in \operatorname{Inn}(L)$. Hence $\gamma_{2} R_{1}$ is a middle inner mapping.
4. It is a consequence of the first property and the fact that any automorphism in a C-loop $L$ is a pseudo-automorphism with companion $x^{2}$ for all $x \in L$.

Lemma 3.3. Let $(L, \cdot)$ be a C-loop. Then:

1. $T\left(x^{-1}\right)=R_{x} T\left(x^{-2}\right) L_{x}^{-1}, T(x)^{2}=R_{x} T\left(x^{-1}\right)^{-1} L_{x}^{-1}$,
2. $T\left(x^{n}\right)=R_{x}^{n-1} T(x) L_{x}^{1-n}, T\left(x^{-n}\right)=R_{x}^{1-n} T\left(x^{-1}\right) L_{x}^{n-1}$ for $n \in \boldsymbol{Z}^{+}$,
3. $R(x, x)=I, L(x, x)=I$.

Proof. 1. For $\gamma_{2} R_{1}$ from Lemma 3.2 we have $y \gamma_{2} R_{1}=x^{2} y x^{-2}=y R_{x^{-2}} L_{x^{2}}=$ $y R_{x}^{-1} R_{x}^{-1} L_{x} L_{x}=y R_{x}^{-1} T\left(x^{-1}\right) L_{x}$. Thus, $\gamma_{2} R_{1}=R_{x}^{-1} T\left(x^{-1}\right) L_{x}$. But $\gamma_{2} R_{1}=T\left(x^{-2}\right)$ is the middle inner mapping, so, $T\left(x^{-2}\right)=R_{x}^{-1} T\left(x^{-1}\right) L_{x}$ implies $T\left(x^{-1}\right)=R_{x} T\left(x^{-2}\right) L_{x}^{-1}$. Therefore $T(x)^{2}=R_{x} L_{x}^{-1} R_{x} L_{x}^{-1}=$ $R_{x}\left(R_{x^{-1}} L_{x^{-1}}^{-1}\right)^{-1} L_{x}^{-1}=R_{x} T\left(x^{-1}\right)^{-1} L_{x}^{-1}$.
2. By induction.
$n=1, T(x)=R_{x}^{1-1} T(x) L_{x}^{1-1}=R_{x^{0}} T(x) L_{x^{0}}=T(x)$ for $x \in L$,
$n=2, T\left(x^{2}\right)=T(x x)=R_{x^{2}} L_{x^{2}}^{-1}=R_{x} R_{x} L_{x}^{-1} L_{x}^{-1}=R_{x} T(x) L_{x}^{-1}$ for $x \in L$,
$n=3, T\left(x^{3}\right)=T\left(x^{2} x\right)=R_{x^{2} x} L_{\left(x^{2} x\right)^{-1}}=R_{x^{2}} R_{x} L_{x^{-1} x^{-2}}=R_{x^{2}} R_{x} L_{x^{-1}} L_{x^{-2}}$ $=R_{x}^{2} T(x) L_{x}^{-2}$ for all $x \in L$.
Let $n=k, T\left(x^{k}\right)=R_{x}^{k-1} T(x) L_{x}^{1-k}$. Then for $n=k+1$ we have

$$
\begin{aligned}
& T\left(x^{k+1}\right)=T\left(x^{k-1} x^{2}\right)=R_{x^{k-1} x^{2}} L_{\left(x^{k-1} x^{2}\right)}^{-1}=R_{x^{k-1} x^{2}} L_{x^{-2} x^{1-k}}= \\
& R_{x^{k-1}} R_{x^{2}} L_{x^{-2}} L_{x^{1-k}}=R_{x^{k-1}} T\left(x^{2}\right) L_{x^{1-k}}=R_{x}^{k-1} R_{x} T(x) L_{x}^{-1} L_{x}^{1-k} \\
& =R_{x}^{k} T(x) L_{x}^{-k}
\end{aligned}
$$

Therefore $T\left(x^{n}\right)=R_{x}^{n-1} T(x) L_{x}^{1-n}$ for all $n \in \boldsymbol{Z}^{+}$. Replacing $x$ by $x^{-1}$ we obtain $T\left(x^{-n}\right)=T\left(\left(x^{-1}\right)^{n}\right)=R_{x^{-1}}^{n-1} T\left(x^{-1}\right) L_{x^{-1}}^{1-n}=R_{x}^{1-n} T\left(x^{-1}\right) L_{x}^{n-1}$. Thus, $T\left(x^{-n}\right)=R_{x}^{1-n} T\left(x^{-1}\right) L_{x}^{n-1}$ for all $n \in \mathbb{Z}^{+}$.
3. $R(x, x)=R_{x}^{2} R_{x}^{-2}=I, L(x, x)=L_{x}^{2} L_{x}^{-2}=I$.

Remark 3.1. Lemma 3.2 gives an example of a bijective mapping which is an automorphism, pseudo-automorphism, semi-automorphism and an inner mapping.

## 4. Relationship between C-loops and Steiner loops

For a loop $(L, \cdot)$, the bijection $J: L \rightarrow L$ is defined by $x J=x^{-1}$. A Steiner loop is a loop satisfying the identities

$$
x^{2}=e, \quad y x \cdot x=y, \quad x y=y x
$$

Theorem 4.1. A $C$-loop $(L, \cdot)$ in which $\left(I, L_{z}^{2}, J L_{z}^{2} J\right.$ or $\left(R_{z}^{2}, I, J R_{z}^{2} J\right)$ lies in $A U T(L)$ is a loop of exponent 4.

Proof. 1. If $\left(I, L_{z}^{2}, J L_{z}^{2} J\right) \in A U T(L)$ for all $z \in L$, then $x \cdot y L_{z}^{2}=(x y) J L_{z}^{2} J$ for all $x, y, z \in L$ implies $x \cdot z^{2} y=x y \cdot z^{-2}$. Whence $z^{2} y \cdot z^{2}=y$. Then $y^{4}=e$ for every $y \in L$.
2. If $\left(R_{z}^{2}, I, J R_{z}^{2} J\right) \in A U T(L)$ for all $z \in L$, then $x R_{z}^{2} \cdot y=(x y) J R_{z}^{2} J$ for all $x, y, z \in L$ implies $\left(x z^{2}\right) \cdot y=\left[(x y)^{-1} z^{2}\right]^{-1}$. Whence $\left(x z^{2}\right) \cdot y=z^{-2}(x y)$, consequently $\left(x z^{2}\right) \cdot y=z^{-2} x \cdot y$. Thus $x z^{2}=z^{-2} x$ which implies $z^{4}=e$ for every $z \in L$.

Theorem 4.2. A $C$-loop $(L, \cdot)$ in which $\left(I, L_{z}^{2}, J L_{z}^{2} J\right)$ and $\left(R_{z}^{2}, I, J R_{z}^{2} J\right)$ lies in $A U T(L)$ is a central square C-loop of exponent 4.

Proof. 1. If $\left(I, L_{z}^{2}, J L_{z}^{2} J\right) \in A U T(L)$ for all $z \in L$, then $x \cdot y L_{z}^{2}=(x y) J L_{z}^{2} J$ for all $x, y, z \in L$ implies $x \cdot z^{2} y=x y \cdot z^{-2}$.
2. If $\left(R_{z}^{2}, I, J R_{z}^{2} J\right) \in A U T(L)$ for all $z \in L$, then $x R_{z}^{2} \cdot y=(x y) J R_{z}^{2} J$ for all $x, y, z \in L$ implies $x z^{2} \cdot y=z^{-2}(x y)$.

Therefore $x \cdot z^{2} y=x z^{2} \cdot y$ if and only if $x y \cdot z^{-2}=z^{-2} \cdot x y$. Putting $t=x y$ we have $t z^{-2}=z^{-2} t$, i.e., $z^{2} t^{-1}=t^{-1} z^{2}$. Whence we conclude that
$z^{2} \in C(L, \cdot)$ for all $z \in L$. Since C-loops are nuclear square (see [26]), we have $z^{2} \in Z(L, \cdot)$. Hence $L$ is a central square C-loop. By Theorem 4.1, $x^{4}=e$.

Corollary 4.1. If $\left(I, L_{z}^{2}, J L_{z}^{2} J\right) \in \operatorname{AUT}(L)$ and $\left(R_{z}^{2}, I, J R_{z}^{2} J\right) \in A U T(L)$ for a C-loop $(L, \cdot)$, then $L$ is flexible, $(x y)^{2}=(y x)^{2}$ for all $x, y \in L$ and $x \mapsto x^{3}$ is an anti-automorphism

Proof. By Theorem 4.2, Lemma 5.1 and Corollary 5.2 of [21].
Theorem 4.3. A central square C-loop of exponent 4 is a group.
Proof. To prove this, it shall be shown that $R(x, y)=I$ for all $x, y \in L$. By Corollary 4.1, for $w \in L$ we get $w R(x, y)=w R_{x} R_{y} R_{x y}^{-1}=(w x) y \cdot(x y)^{-1}=$ $(w x)\left(x^{2} y x^{2}\right) \cdot(x y)^{-1}=\left(w x^{3}\right)\left(y x^{2}\right) \cdot(x y)^{-1}=\left(w^{2}\left(w^{3} x^{3}\right)\right)\left(y x^{2}\right) \cdot(x y)^{-1}=$ $\left(w^{2}(x w)^{3}\right)\left(y x^{2}\right) \cdot(x y)^{-1}=w^{2}(x w)^{3} \cdot\left(y x^{2}\right)(x y)^{-1}=w^{2}(x w)^{3} \cdot\left[y \cdot x^{2}(x y)^{-1}\right]=$ $w^{2}(x w)^{3} \cdot\left[y \cdot x^{2}\left(y^{-1} x^{-1}\right)\right]=w^{2}(x w)^{3} \cdot\left[y\left(y^{-1} x^{-1} \cdot x^{2}\right)\right]=w^{2}(x w)^{3} \cdot\left[y\left(y^{-1} x\right)\right]=$ $w^{2}(x w)^{3} \cdot x=w^{2}\left(w^{3} x^{3}\right) \cdot x=w^{2} \cdot\left(w^{3} x^{3}\right) x=w^{2} \cdot\left(w^{3} x^{-1}\right) x=w^{2} w^{3}=$ $w^{5}=w$. So, $R(x, y)=I$, i.e., $R_{x} R_{y} R_{x y}^{-1}=I$. Thus $R_{x} R_{y}=R_{x y}$ and $z R_{x} R_{y}=z R_{x y}$. So, $z x \cdot y=z \cdot x y$. Therefore $L$ is a group.

Corollary 4.2. A C-loop $(L, \cdot)$ in which for all $z \in L\left(I, L_{z}^{2}, J L_{z}^{2} J\right)$ and $\left(R_{z}^{2}, I, J R_{z}^{2} J\right)$ are in $\operatorname{AUT}(L)$ is a group.

Proof. This follows from Theorem 4.2 and Theorem 4.3.
Remark 4.1. Central square C-loops of exponent 4 are A-loops.
Theorem 4.4. A C-loop is a central square loop if and only if $\gamma_{2} R_{1}=I$.
Proof. $\gamma_{2} R_{1}=I \longleftrightarrow T\left(x^{-2}\right)=I$ for all $x \in L \longleftrightarrow R_{x^{-2}} L_{x^{2}}=I \longleftrightarrow$ $y x^{2}=x^{2} y \longleftrightarrow L$ is central square.

Theorem 4.5. Let $L$ be a C-loop such that the mapping $x \mapsto T(x)$ is a bijection, then $L$ is of exponent 2 if and only if $\gamma_{2} R_{1}=I$.

Proof. Indeed, $\gamma_{2} R_{1}=I \longleftrightarrow T\left(x^{-2}\right)=I$ for all $x \in L \longleftrightarrow T\left(x^{-2}\right)=$ $I=R_{x}^{-1} T\left(x^{-1}\right) L_{x} \longleftrightarrow T\left(x^{-1}\right)=T(x) \longleftrightarrow x^{-1}=x$. Since $x \mapsto T(x)$ is a bijection $L$ is a loop of exponent 2 .

Corollary 4.3. A C-loop in which $x \mapsto T(x)$ is a bijection is a loop of exponent 2 if and only if it is central square.

Proof. By Theorem 4.4 and Theorem 4.5.
Corollary 4.4. A central square C-loop in which the map $x \mapsto T(x)$ is a bijection is a Steiner loop.

Proof. By the converse of Corollary 4.3, a C-loop in which $x \mapsto T(x)$ is a bijection, is of exponent 2 if it is central square. By the result of [26], an inverse property loop of exponent 2 is a Steiner loop. By the fact that C-loops are inverse property loops [26], it is a Steiner loop.

Corollary 4.5. A C-loop $(L, \cdot)$ in which $x \mapsto T(x)$ is a bijection and $\left(I, L_{z}^{2}, J L_{z}^{2} J\right),\left(R_{z}^{2}, I, J R_{z}^{2} J\right)$ are in $\operatorname{AUT}(L)$ for every $z \in L$, is a Steiner loop of exponent 4.

Proof. According to Theorem 4.2, $L$ is a central square loop. Since $x \mapsto$ $T(x)$ is a bijection, by Corollary 4.4, $L$ is a Steiner loop. By Theorem 4.1, it has a an exponent of 4 .

Corollary 4.6. A C-loop $L$ in which the mapping $x \mapsto T(x)$ is a bijection is a Steiner loop if and only if $L$ is a central square $C$-loop.

Proof. A Steiner loop $L$ is a C-loop [26]. Steiner loops are loops of exponent two, hence by Corollary 4.3, $L$ is central square since in $L$, the mapping $x \mapsto T(x)$ is a bijection. Conversely, by Corollary 4.3, a central square C-loop $L$ in which the mapping $x \mapsto T(x)$ is a bijection is a loop of of exponent two. The fact that an inverse property loop of exponent two is a Steiner loop [26], completes the proof.

### 4.1. Flexibility in C-loops

Lemma 4.1. A C-loop is flexible if the mapping $x \mapsto x^{2}$ is onto.
Proof. Let $L$ be a C-loop. Then $y x^{2} \cdot y=y \cdot x^{2} y$ for all $x, y \in L$. Thus, $L$ is square flexible, hence by [12], it is flexible since the mapping $x \mapsto x^{2}$ is onto.

Theorem 4.6. A C-loop $L$ is flexible if $\left(I, L_{z}^{2}, J L_{z}^{2} J\right)$ and $\left(R_{z}^{2}, I, J R_{z}^{2} J\right)$ are in $\operatorname{AUT}(L)$ for all $z \in L$ and the middle inner mappings are of order 2 .

Proof. By Lemma 3.3, for every $x \in L$ we have $T(x)^{2}=R_{x} T\left(x^{-1}\right)^{-1} L_{x}^{-1}=$ $R_{x}\left(R_{x} T\left(x^{-2}\right) L_{x}^{-1}\right)^{-1} L_{x}^{-1}=R_{x}\left(L_{x}\left(R_{x} T\left(x^{-2}\right)\right)^{-1}\right) L_{x}^{-1}=R_{x}\left(L_{x} T\left(x^{-2}\right)^{-1}\right.$ $\left.R_{x}^{-1}\right) L_{x}^{-1}=R_{x} L_{x} T\left(x^{-2}\right)^{-1} R_{x}^{-1} L_{x}^{-1}=R_{x} L_{x} T\left(x^{-2}\right)^{-1}\left(L_{x} R_{x}\right)^{-1}$. Therefore
$T(x)^{2}=R_{x} L_{x} T\left(x^{-2}\right)^{-1}\left(L_{x} R_{x}\right)^{-1} \longleftrightarrow T(x)^{2} L_{x} R_{x}=R_{x} L_{x} T\left(x^{-2}\right)^{-1}=$ $R_{x} L_{x}\left(\gamma_{2} R_{1}\right)^{-1}=R_{x} L_{x} \gamma_{1} R_{2} \longleftrightarrow T(x)^{2} L_{x} R_{x}=R_{x} L_{x} \gamma_{1} R_{2}$. If $|T(x)|=2$, $T(x)^{2}=I$ and if $\gamma_{1} R_{2}=I \longleftrightarrow L$ is central square, then $L_{x} R_{x}=R_{x} L_{x} \longleftrightarrow$ $x y \cdot x=x \cdot y x$ is a flexible loop.

Philips and Vojtěchovský [26] studied the close relationship between Cloops and Steiner loops. In [23], it is shown that Steiner loops are exactly commutative inverse property loops of exponent 2 . But in [26], this fact was improved, so that commutativity is not a sufficient condition for an inverse property loop of exponent 2 to be a Steiner loop. So they said 'Steiner loops are exactly inverse property loops of exponent 2 '. This result is general for inverse property loops among which are C-loops. They also proved that Steiner loops are C-loops.

The flexibility is possible in a C-loop if the loop is commutative or diassociative [23]. But C-loops naturally do not even satisfy the latter. Apart from the condition stated in Lemma 4.1, Theorem 4.6 when compared with Corollary 5.2 of [21] shows that some middle inner-mappings do not need to be of exponent of 2 for a C-loop to be flexible.

## 5. Unique non-identity commutator and associator

Lemma 5.1. If $s$ is a unique non-identity commutator in a $C$-loop $L$, then $|s|=2, s \in C(L)$ and $s \in Z\left(L^{2}\right)$.

Proof. $x y=(y x)(x, y) \longleftrightarrow(x, y)=(y x)^{-1}(x y)=\left(x^{-1} y^{-1}\right)(x y)$. Therefore $(x, y)^{-1}=\left[\left(x^{-1} y^{-1}\right)(x y)\right]^{-1}=(x y)^{-1}\left(x^{-1} y^{-1}\right)^{-1}=\left(y^{-1} x^{-1}\right)(y x)=$ $(y, x)$. Thus, $s^{-1}=s$ or $s^{-1}=e$ implies $s^{2}=e$ or $s=e$. So, $s^{2}=e$.

If $x s \neq s x$, then $x s=(s x) s$ implies $x=s x$, whence $s=e$. So, $x s=s x$, i.e., $s \in C(L)$. Hence, $s \in Z\left(L^{2}\right)$.

Lemma 5.2. If $s$ is a unique non-identity associator in a $C$-loop $L$, then $s \in N(L)$.

Proof. If $(x y) s \neq x(y s)$, then $(x y) s=x(y s) \cdot s$ implies $x y=x \cdot y s$. Whence $y=y s$, i.e., $s=e . \operatorname{So},(x y) s=x(y s)$, that is, $s \in N(L)$.

Lemma 5.3. If a $C$-loop $(L, \cdot)$ has a unique non-identity commutator associator $s$, then $s$ is a central element of order 2 .

Proof. We shall keep in mind that $L$ as a C-loop has the inverse property. $s \in(L, L)$ implies $s^{-1} \in(L, L)$, whence $s^{-1}=s$. Since $s^{-1} \neq e$ if and only if $s \neq e$, we have $s^{2}=e$. Let $x s \neq s x$ for some $x, y \in L$. Then $x s=(s x) s$ implies $x=s x$, i.e., $s=e$, which is a contradiction. Thus, $s \in C(L)$. If $(x y) s \neq x(y s)$ for some $x, y \in L$, then $(x y) s=(x \cdot y s) s$ implies $x y=x \cdot y s$. Thus $y=y s$, i.e, $s=e$, which is a contradiction. So, $s \in N(L)$. Therefore $s \in C(L), s \in N(L)$ implies $s \in Z(L)$.

Remark 5.1. The result of Lemma 5.3 is similar to the result proved in [16] for Moufang loops.

Lemma 5.4. In $L C(R C)$-loops with a unique non-identity square s is $|s|=$ $2,|x|=4$ or $|x|=2, s \in N_{\lambda}$ or $s \in N_{\rho}$ and $s \in N_{\mu}$.

Proof. For all $x \in L$ we have $x^{2}=s$. Since $s^{2}=s$ implies $s^{-1} s^{2}=s^{-1} s$ or $s^{2} s^{-1}=s s^{-1}$, so $s=e$. This is a contradiction, thus $s^{2}=e$ if and only if $|s|=2$. Moreover, $x^{2}=s$ implies $x^{4}=x^{2} x^{2}=s^{2}=e$. Therefore $x^{4}=e$ or $x^{2}=e$. In any LC-loop, $x^{2} \in N_{\lambda}, N_{\mu}$, thus $s \in N_{\lambda}, N_{\mu}$. In an RC-loop, $x^{2} \in N_{\rho}, N_{\mu}$, thus $s \in N_{\rho}, N_{\mu}$.

Lemma 5.5. An $L C(R C)$-loop $L$ has a unique non-identity square $s$ if and only if $J=R_{s}^{-1}=R_{s^{-1}}^{-1}$ or $J=I\left(\right.$ resp. $J=L_{s}^{-1}=L_{s^{-1}}^{-1}$ or $\left.J=I\right)$.

Proof. Let $L$ be a RC-loop. Then $x^{2}=s \longleftrightarrow x^{2} x^{-1}=s x^{-1} \longleftrightarrow x=$ $s x^{-1} \longleftrightarrow x=x J L_{s} \longleftrightarrow I=J L_{s} \longleftrightarrow J=L_{s}^{-1}=L_{s^{-1}}^{-1}$. Similarly, $x^{2}=e \longleftrightarrow x=x^{-1} \longleftrightarrow x=x J \longleftrightarrow J=I$.

For LC-loops the proof is analogous.
Theorem 5.1. For any L.I.P. (R.I.P.) RC(LC)-loop ( $L, \cdot$ ) with a unique non-identity square $s$,

1. $s \in Z(L, \cdot)$, i.e., $L$ is centrum square,
2. $J=L_{s} \quad\left(\right.$ resp. $\left.J=R_{s}\right)$,
3. $x^{2} y^{2} \neq(x y)^{2} \neq y^{2} x^{2}$, i.e., $x \mapsto x^{2}$ is neither an automorphism nor an anti-automorphism,
4. $(a, b, c)=(b c \cdot a)(a b \cdot c)$,
(a) $a b=a^{-1} b^{-1}$ if and only if $(J, J, I) \in \operatorname{AUT}(L)$,
(b) $(a, b, a)=(b s)(a b \cdot a)$ or $(a, b, a)=b(a b \cdot a)$,
5. L is a group or Steiner loop,
6. If $L$ is a non-commutative $C$-loop, then $s$ is its unique non-identity commutator.

Proof. 1. $x^{2}=s$ implies $x=s x^{-1}$, whence $x^{-1}=s^{-1} x$. This, by Lemma 2.1 from [1], gives $x^{-1}=\left(s x^{-1}\right)^{-1}=\left(x^{-1}\right)^{-1} s^{-1}=x s^{-1}$. Thus, $x^{-1}=s^{-1} x=x s^{-1}$, i.e., $s x=x s$. So, $s \in Z(L, \cdot)$.
2. This follows from Lemma 5.5.
3. If $(x y)^{2}=x^{2} y^{2}$ or $(x y)^{2}=y^{2} x^{2}$, then $s=s^{2}$ implies $s=e$ which is a contradiction.
4. $(a, b, c)=[a(b c)]^{-1} \cdot(a b) c=(b c)^{-1} a^{-1} \cdot(a b) c=\left(c^{-1} b^{-1}\right) a^{-1} \cdot(a b \cdot c)=$ $\left[s^{-1}(b c)\right]\left(s^{-1} a\right) \cdot(a b \cdot c)=\left(b c \cdot s^{-1}\right)\left(s^{-1} a\right) \cdot(a b \cdot c)=\left(b c s^{-2} \cdot a\right)(a b \cdot c)=$ $(b c \cdot a)(a b \cdot c)$. So, $(a, b, c)=(b c \cdot a)(a b \cdot c)$.
$4 a$. The above for $c=e$ gives $(a, b, e)=(b a)(a b)=e$, whence $a b=(b a)^{-1}=$ $a^{-1} b^{-1}$. So, $(J, J, I) \in A U T(L)$.
4b. For $c=a$ we have $(a, b, a)=(b a \cdot a)(a b \cdot a)=\left(b a^{2}\right)(a b \cdot a)=(b s)(a b \cdot a)$. Thus $(a, b, a)=(b s)(a b \cdot a)$ or $(a, b, a)=b(a b \cdot a)$.
5. This follows from Lemma 5.4.
6. $(x, y)=x^{-1} y^{-1} \cdot x y=\left(x^{-1} y^{-1}\right)\left(x y^{-1} \cdot y^{2}\right)=\left(\left(x^{-1} y^{-1}\right)\left(x y^{-1}\right) \cdot y^{2}=\right.$ $\left[x^{-2}\left(x y^{-1}\right) \cdot\left(x y^{-1}\right)\right] y^{2}=x^{-2}\left[\left(x y^{-1}\right)\left(x y^{-1}\right)\right] y^{2}=e$ or $(x, y)=s$. Thus, $L$ is either commutative or $s$ is its unique non-identity commutator.

For $(x, s)=x^{-1} s^{-1} \cdot x s=s$ we have $x^{-1} R_{s} \cdot x R_{s}=s$, whence $x J^{2} \cdot x^{-1}=$ $s$. Thus $x x^{-1}=s$, i.e., $s=e$, which is a contradiction. So. $(x, s)=e$ implies $s \in C(L, \cdot)$.

Corollary 5.1. A C-loop with a unique non-trivial square is a group.
Proof. By Lemma 5.4 and Theorem 5.1, it is central square of exponent 4. By Theorem 4.3, it is a group.

Remark 5.2. A C-loop with a unique non-trivial square is an A-loop.
Theorem 5.2. Let $(G, \cdot)$ and $(H, \circ)$ be two distinct loop such that the triple $\alpha=(A, B, C)$ is an isotopism of $G$ onto $H$.

1. If $G$ is a central square $C$-loop of exponent 4 , then $H$ is a $C$-loop and an A-loop.
2. If $G$ is a C-loop with a unique non-identity square, then $H$ is a $C$-loop and an A-loop.

Proof. 1. By Theorem 4.3, $G$ is a group and since groups are G-loops, $H$ is a group, i.e., it is a C-loop and an A-loop.
2. By Corollary 5.1.

Remark 5.3. Some results for isotopes of central loops of the type $(A, B, B)$ and $(A, B, A)$ are obtained in [18].

Corollary 5.2. Let $(G, \cdot)$ and $(H, \circ)$ be distinct loops. If the triple $(A, B, C)$ is an isotopism of $G$ onto $H$ such that for every $z \in G\left(I, L_{z}^{2}, J L_{z}^{2} J\right)$ and $\left(R_{z}^{2}, I, J R_{z}^{2} J\right)$ are in $\operatorname{AUT}(G, \cdot)$, then $H$ is a C-loop and an $A$-loop.

Proof. It follows from Theorem 4.2 and Theorem 5.2.
Theorem 5.3. An isotopism $(A, A, C)$ saves the property "unique nonidentity square".

Proof. Let $(A, A, C):(G, \cdot) \rightarrow(H, \circ)$, where $G$ and $H$ are two distinct loops, be an isotopism. Then $x A \circ y A=(x \cdot y) C$. For $y=x$ we have $x A \circ x A=(x A)^{2}=(x \cdot x) C=x^{2} C$. If $s$ is the unique non-identity square in $G$, i.e $x^{2}=s$ or $x^{2}=e$ for all $x \in G$ then $s^{\prime}=s C=(x A)^{2}=y^{\prime 2}$ or $y^{\prime 2}=(x A)^{2}=x^{2} C=e C=e^{\prime}$ for all $y^{\prime} \in H$ with $e^{\prime}$ as the identity element in $H$. So, $s^{\prime}$ is the unique non-identity square element in $H$.

Corollary 5.3. Central loops with unique non-identity square are isotopic invariant.

## 6. Cross inverse property in central loops

According to [5], the W.I.P. is a generalization of the C.I.P. The latter was introduced and studied by R. Artzy [3] and [4], but from the formal point of view this property was introduced by J. M. Osborn [22]. Huthnance Jr. [17], proved that the holomorph of a W.I.P.L. is a W.I.P.L. A loop property is called universal (or universal relative to a given property) if every loop isotope of this loop is a loop with this property. A universal W.I.P.L. is called an Osborn loop. Huthnance Jr. [17] investigated the structure of some holomorph of Osborn loops. Basarab [6] studied Osborn loops which are G-loops.

Theorem 6.1. An $L C(R C)$-loop of exponent 3 is centrum square if and only if it is a C.I.P.L.

Proof. Let $L$ be a LC-loop. Then $x^{2} y=y x^{2} \longleftrightarrow x^{-1} y=y x^{-1} \longleftrightarrow$ $x\left(x^{-1} y\right)=x\left(y x^{-1}\right) \longleftrightarrow y=x\left(y x^{-1}\right)$, which holds if and only if the C.I.P. holds in $L$.

For RC-loops the proof is analogous.
Corollary 6.1. If $L$ is a centrum square $L C(R C)$-loop of exponent 3 , then

1. L has the A.I.P. and A.A.I.P.,
2. L has the W.I.P.,
3. $N=N_{\lambda}=N_{\rho}=N_{\mu}$,
4. $n \in N$ implies $n \in Z(L)$,
5. $L$ is a commutative group.

Proof. 1. By Theorem 6.1, $L$ is a C.I.P.L. According to [4] and [5], a C.I.P.L. has the A.I.P. Thus, the first part is true. The second part follows from the fact that $x^{2}=x^{-1}$.
2. This follows from the the fact that W.I.P. is a generalization of C.I.P. [23].
3. and 4. follows from [5] and [4]. The last statement is obvious.

Lemma 6.1. Any $L C(R C, C)$-loop of exponent 3 is a group.
Corollary 6.2. A central square C-loop of exponent 3 has the W.I.P. and C.I.P. and it a commutative group.

The fact that central loops of exponent 3 are groups it will be interesting to study non-commutative central loops of exponent 3 with the C.I.P. since there exist groups that do not have the C.I.P. From Theorem 6.1, it follows that the study of $\mathrm{LC}(\mathrm{RC})$-loops of exponent 3 with C.I.P. is equivalent to the study of centrum square $\mathrm{LC}(\mathrm{RC})$-loops of exponent 3 .

The existence of central loops of exponent 3 can be deduced from [15], [26] and [27]. According to [26] and [27], the order of every element in a finite $\mathrm{LC}(\mathrm{RC})$-loop divides the order of the loop. Since $|x|=3$ for all $x \in L$, then
$|L|=2 m, m \geqslant 3$ if $L$ is a non-left (right) Bol LC(RC)-loop, or
$|L|=4 k, k>2$ if $L$ is a non-Moufang but both left (right)-Bol and LC(RC)-loop.
The possible orders of finite RC-loops were calculated in [27].

### 6.1. Osborn central-loops

Theorem 6.2. An $L C(R C)$-loop has the R.I.P. (L.I.P.) if and only if has the W.I.P.

Proof. Let $(L, \cdot)$ be a LC-loop with the W.I.P. Then for all $x, y \in L$, $y(x y)^{\rho}=x^{\rho}$. Let $x y=z$, then $x^{\lambda}(x y)=x^{\lambda} z$ implies $y=x^{\lambda} z$, thus $\left(x^{\lambda} z\right) z^{\rho}=x^{\rho}$ implies $\left(x^{-1} z\right) z^{\rho}=x^{-1}$. Replacing $x^{-1}$ by $x$, we obtain $(x z) z^{\rho}=x$. So, $L$ has the R.I.P.

Conversely, if $L$ has the I.P., then $y(x y)^{\rho}=y(x y)^{-1}=y\left(y^{-1} x^{-1}\right)=$ $x^{-1}=x^{\rho}$ hence it has the W. I. P. Let $L$ be a RC-loop with the W.I.P. Then for all $x, y \in L, y(x y)^{\rho}=x^{\rho}$ if and only if $(x y)^{\lambda} \cdot x=y^{\lambda}$. Let $x y=z$, then $(x y) y^{\rho}=z y^{\rho}$ implies $x=z y^{\rho}$. Thus, $z^{\lambda}\left(z y^{\rho}\right)=y^{\lambda}$ implies $z^{\lambda}\left(z y^{-1}\right)=y^{-1}$. Replacing $y^{-1}$ by $y$, we get $z^{\lambda}(z y)=y$. Thus, $L$ has the L.I.P.

Corollary 6.3. Let $(L, \cdot)$ be an $L C(R C)$-loop with R.I.P. (L.I.P.). Then

1. $N(L)=N_{\lambda}(L)=N_{\rho}(L)=N_{\mu}(L)$,
2. $\left(I, R_{x^{2}}, R_{x^{2}}\right) \in A U T(L)\left(\right.$ resp. $\left(L_{x^{2}}, I, L_{x^{2}}\right) \in A U T(L)$,
3. $\left(L_{x}^{2}, R_{x^{2}}, R_{x^{2}} L_{x}^{2}\right) \in \operatorname{AUT}(L) \quad\left(\right.$ resp. $\left(L_{x^{2}}, R_{x}^{2}, L_{x^{2}} R_{x}^{2}\right) \in \operatorname{AUT}(L)$.

Proof. By Theorem 6.2, L has the W.I.P. According to [22], in a W.I.P.L., the three nuclei coincide, so the firs statement is true. Thus for an LCloop, $x^{2} \in N_{\rho}$ and for an RC-loop, $x^{2} \in N_{\lambda}$. Hence for an LC-loop $L,\left(L_{x}^{2}, I, L_{x}^{2}\right),\left(I, R_{x^{2}}, R_{x^{2}}\right) \in \operatorname{AUT}(L)$ implies that $\left(L_{x}^{2}, R_{x^{2}}, L_{x}^{2} R_{x^{2}}\right)=$ $\left(L_{x}^{2}, R_{x^{2}}, R_{x^{2}} L_{x}^{2}\right) \in \operatorname{AUT}(L)$. For an RC-loop $L,\left(I, R_{x}^{2}, R_{x}^{2}\right),\left(L_{x^{2}}, I, L_{x^{2}}\right) \in$ $\operatorname{AUT}(L)$ implies $\left(L_{x^{2}}, R_{x}^{2}, R_{x}^{2} L_{x^{2}}\right)=\left(L_{x^{2}}, R_{x}^{2}, L_{x^{2}} R_{x}^{2}\right) \in \operatorname{AUT}(L)$. So, the last two statement are true, too.

Remark 6.1. Corollary 6.3 is true for left (right) Bol loops (i.e., LB(RB)loops). It follows from the fact that a $\mathrm{RB}(\mathrm{LB})-\mathrm{loop}$ has the L.I.P. (R.I.P.) if and only if it is a Moufang loop [23], which is obviously a W.I.P.L. [19].

Theorem 6.3. An $L C(R C)$-loop $L$ is a C-loop if and only if one of the following equivalent statements holds:

1. L has the R.I.P. (L.I.P.),
2. L has the R.A.P. (L.A.P.),
3. $L$ is a $R C(L C)$-loop,
4. $L$ has the A.A.I.P. (i.e., $\left.(x y)^{-1}=y^{-1} x^{-1}\right)$,
5. L has the W.I.P.

Proof. A C-loop satisfies 1 and 2. Conversely, if $L$ is an LC-loop, then $(x \cdot x y) z=x(x \cdot y z)$, whence $[(x \cdot x y) z]^{-1}=[x(x \cdot y z)]^{-1}$. Thus $z^{-1}(x \cdot x y)^{-1}=$ $(x \cdot y z)^{-1} x^{-1}$ and consequently $z^{-1}\left((x y)^{-1} \cdot x^{-1}\right)=\left((y z)^{-1} \cdot x^{-1}\right) x^{-1}$, i.e., $z^{-1}\left(y^{-1} x^{-1} \cdot x^{-1}\right)=\left(z^{-1} y^{-1} \cdot x^{-1}\right) x^{-1}$, which means that $z(y x \cdot x)=(z y \cdot x) x$ for all $x, y, z \in L$. So, a RC-loop. Hence, $L$ is a C-loop.

If $L$ is an LC-loop, then according to [26], $x \cdot(y \cdot y z)=(x \cdot y y) z$ for all $x, y, z \in L$, while $L$ is an RC-loop if and only if $(z y \cdot y) x=z(y y \cdot x)$ for all $x, y, z \in L$. Thus $x \cdot(y \cdot y z)=(x \cdot y y) z$, or equivalently $x \cdot z L_{y}^{2}=x R_{y^{2}} \cdot z$. So, $\left(R_{y^{2}}, L_{y}^{-2}, I\right) \in A U T(L)$ for all $y \in L$. For $(z y \cdot y) x=z(y y \cdot x)$ we have $z R^{2} \cdot x=z \cdot x L_{y^{2}}$, i.e., $\left(R_{y}^{2}, L_{y^{2}}^{-1}, I\right) \in A U T(L)$ for all $y \in L$.

If $L$ has the right (left) alternative property, $\left(R_{y}^{2}, L_{y}^{-2}, I\right) \in A U T(L)$ for all $y \in L$ if and only if $L$ is a C-loop.
3. This is shown in [15].
4. This is equivalent to 1 . Indeed, if $L$ has the L.I.P. (R.I.P.), then $L$ has the R.I.P. (L.I.P.). so, $L$ has the A.A.I.P. Conversely, if L.I.P. holds, then for $z=x y$, we have $y=x^{-1} z$ which gives $z^{-1}=\left(x^{-1} z\right)^{-1} x^{-1}$, whence $z^{-1}=\left(z^{-1} x\right) x^{-1}$. So, $z=(z x) x^{-1}$.

Similarly, if $L$ has the R.I.P. (L.I.P.) then $L$ has the L.I.P. (R.I.P.), i.e., it has the A.A.I.P. Conversely, if R.I.P. holds, then for $z=x y$, we have $x=z y^{-1}$. Thus, $z^{-1}=y^{-1}\left(z y^{-1}\right)^{-1}=y^{-1}\left(y z^{-1}\right)$, which proves the L.I.P.
5. This follows from 1 and Theorem 6.2.

Theorem 6.4. (cf. [19]) The following equivalent conditions define an Osborn loop $(L, \cdot)$.

1. $x(y z \cdot x)=\left(x \cdot y E_{x}\right) \cdot z x$,

2 . $(x \cdot y z) x=x y \cdot\left(z E_{x}^{-1} \cdot x\right)$,
3. $\left(A_{x}, R_{x}, R_{x} L_{x}\right) \in A U T(L)$,
4. $\left(L_{x}, B_{x}, L_{x} R_{x}\right) \in A U T(L)$,
where $A_{x}=E_{x} L_{x}, B_{x}=E_{x}^{-1} R_{x}$ and $E_{x}=R_{x} L_{x} R_{x}^{-1} L_{x}^{-1}$.
Theorem 6.5. If a $R C(L C)$-loop has the L.I.P. (R.I.P.), then it is an Osborn loop if every its element is a square.

Proof. Let $L$ be an RC-loop with L.I.P. Then, by Theorem 6.2, $L$ has the W.I.P. Therefore $\left(A_{x^{2}}, I, L_{x^{2}}\right) \in A U T(L) \longleftrightarrow y A_{x^{2}} \cdot z=(y z) L_{x^{2}}$. But
$(y z) L_{x^{2}}=y E_{x^{2}} L_{x^{2}} \cdot z=y R_{x^{2}} L_{x^{2}} R_{x^{2}}^{-1} L_{x^{2}}^{-1} L_{x^{2}} \cdot z=y R_{x^{2}} L_{x^{2}} R_{x^{2}}^{-1} \cdot z=$ $y R_{x}^{2} L_{x^{2}} R_{x^{2}}^{-1} \cdot z=y L_{x^{2}} R_{x}^{2} R_{x^{2}}^{-1} \cdot z=y L_{x^{2}} \cdot z$. This is equivalent to the fact that $\left(L_{x^{2}}, I, L_{x^{2}}\right) \in \operatorname{AUT}(L)$ for all $x \in L$, which is true by Corollary 6.3.

Thus, $\left(I, R_{x}^{2}, R_{x}^{2}\right)\left(A_{x^{2}}, I, L_{x^{2}}\right)=\left(A_{x^{2}}, R_{x^{2}}, R_{x^{2}} L_{x^{2}}\right) \in \operatorname{AUT}(L)$. Using Theorem 6.4, we see that $L$ is an Osborn loop if every element in $L$ is a square.

Now, let $L$ be an LC-loop. If $L$ has the R.I.P., then, by Theorem 6.2, $L$ has the W.I.P. So, $\left(I, B_{x^{2}}, R_{x^{2}}\right) \in \operatorname{AUT}(L)$ if and only if $y \cdot z B_{x^{2}}=$ $(y z) R_{x^{2}}$. But $(y z) R_{x^{2}}=y \cdot z E_{x^{2}}^{-1} R_{x^{2}}=y \cdot z\left(R_{x^{2}} L_{x^{2}} R_{x^{2}}^{-1} L_{x^{2}}^{-1}\right)^{-1} R_{x^{2}}=$ $y \cdot z L_{x^{2}} R_{x^{2}} L_{x^{2}}^{-1} R_{x^{2}}^{-1} R_{x^{2}}=y \cdot z L_{x^{2}} R_{x^{2}} L_{x^{2}}^{-1}=y \cdot z R_{x^{2}} L_{x}^{2} L_{x^{2}}^{-1}=y \cdot z R_{x^{2}}$. This is equivalent to the fact that $\left(I, R_{x^{2}}, R_{x^{2}}\right) \in \operatorname{AUT}(L)$ for all $x \in L$, which is true by Corollary 6.3.

Thus, $\left(L_{x}^{2}, I, L_{x}^{2}\right)\left(I, B_{x^{2}}, R_{x^{2}}\right)=\left(L_{x^{2}}, B_{x^{2}}, L_{x^{2}} R_{x^{2}}\right) \in \operatorname{AUT}(L)$. Whence, as in previous case, we conclude that $L$ is an Osborn loop if every element in $L$ is a square.

Corollary 6.4. An $L C(R C)$-loop with R.I.P. (L.I.P.) is an Osborn loop if every its element is a square. Hence, this loop is a group.

Proof. This follows from Theorem 6.5. The last conclusion is as a consequence of the fact that $x^{2} \in N(L)$.

Corollary 6.5. A C-loop is an Osborn loop if every its element is a square. Hence, this loop is a group.

Question. Does there exist a C-loop which is an Osborn loop but it is non-associative, non Moufang and non-conjugacy closed?
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# On reconstructing reducible n-ary quasigroups and switching subquasigroups 

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#### Abstract

(1) We prove that, provided $n \geqslant 4$, a permutably reducible $n$-ary quasigroup is uniquely specified by its values on the $n$-ples containing zero. (2) We observe that for each $n, k \geqslant 2$ and $r \leqslant\lfloor k / 2\rfloor$ there exists a reducible $n$ ary quasigroup of order $k$ with an $n$-ary subquasigroup of order $r$. As corollaries, we have the following: (3) For each $k \geqslant 4$ and $n \geqslant 3$ we can construct a permutably irreducible $n$-ary quasigroup of order $k$. (4) The number of $n$-ary quasigroups of order $k>3$ has double-exponential growth as $n \rightarrow \infty$; it is greater than $\exp \exp (n \ln \lfloor k / 3\rfloor)$ if $k \geqslant 6$, and $\exp \exp \left(\frac{\ln 3}{3} n-\right.$ $0.44)$ if $k=5$.


## 1. Introduction

An $n$-ary operation $f: \Sigma^{n} \rightarrow \Sigma$, where $\Sigma$ is a nonempty set, is called an $n$-ary quasigroup or $n$-quasigroup (of order $|\Sigma|$ ) iff in the equality $z_{0}=$ $f\left(z_{1}, \ldots, z_{n}\right)$ knowledge of any $n$ elements of $z_{0}, z_{1}, \ldots, z_{n}$ uniquely specifies the remaining one [2].

An $n$-ary quasigroup $f$ is permutably reducible iff

$$
f\left(x_{1}, \ldots, x_{n}\right)=h\left(g\left(x_{\sigma(1)}, \ldots, x_{\sigma(k)}\right), x_{\sigma(k+1)}, \ldots, x_{\sigma(n)}\right)
$$

where $h$ and $g$ are $(n-k+1)$-ary and $k$-ary quasigroups, $\sigma$ is a permutation, and $1<k<n$. In what follows we omit the word "permutably" because we consider only such type of reducibility.

[^5]We will use the following standard notation: $x_{i}^{j}$ denotes $x_{i}, x_{i+1}, \ldots, x_{j}$.
In Section 2 we show that a reducible $n$-quasigroup can be reconstructed by its values on so-called 'shell'. 'Shell' means the set of variable values with at least one zero.

In Section 3 we consider the questions of imbedding $n$-quasigroups of order $r$ into $n$-quasigroups of order $k \geqslant 2 r$.

In Section 4 we prove that for all $n \geqslant 3$ and $k \geqslant 4$ there exists an irreducible $n$-quasigroup of order $k$. Before, the question of existence of irreducible $n$-quasigroups was considered by Belousov and Sandik [3] ( $n=3$, $k=4)$, Frenkin [5] $(n \geqslant 3, k=4)$, Borisenko [4] ( $n \geqslant 3$, composite finite $k)$, Akivis and Goldberg [7, 8, 1] (local differentiable $n$-quasigroups), Glukhov [6] $(n \geqslant 3$, infinite $k)$.

In Sections 5 and 6 we prove the double-exponential $(\exp \exp (c(k) n))$ lower bound on the number $|Q(n, k)|$ of $n$-quasigroups of finite order $k \geqslant 4$. Before, the following asymptotic results on the number of $n$-quasigroups of fixed finite order $k$ were known:

- $|Q(n, 2)|=2$.
- $|Q(n, 3)|=3 \cdot 2^{n}$, see, e.g., [13]; a simple way to realize this fact is to show by induction that the values on the shell uniquely specify an $n$-quasigroup of order 3 .
- $|Q(n, 4)|=3^{n+1} 2^{2^{n}+1}(1+o(1))[15,11]$.

Note that by the "number of $n$-quasigroups" we mean the number of mutually different $n$-ary quasigroup operations $\Sigma^{n} \rightarrow \Sigma$ for a fixed $\Sigma,|\Sigma|=k$ (sometimes, by this phrase one means the number of isomorphism classes). As we will see, for every $k \geqslant 4$ there is $c(k)>0$ such that $|Q(n, k)| \geqslant 2^{2^{c(k) n}}$. More accurately (Theorem 3), if $k=5$ then $|Q(n, 5)| \geqslant 2^{3^{n / 3-c o n s t}}$; for even $k$ we have $|Q(n, k)| \geqslant 2^{(k / 2)^{n}} ;$ for $k \equiv 0 \bmod 3$ we have $|Q(n, k)| \geqslant 2^{n(k / 3)^{n}}$; and for every $k$ we have $|Q(n, k)| \geqslant 2^{1.5\lfloor k / 3\rfloor^{n}}$. Observe that dividing by the number (e.g., $\left.(n+1)!(k!)^{n}\right)$ of any natural equivalences (isomorphism, isotopism, paratopism,...) does not affect these values notably; so, for the number of equivalence classes almost the same bounds are valid. For the known exact numbers of $n$-quasigroups of order $k$ with small values of $n$ and $k$, as well as the numbers of equivalence classes for different equivalences, see the recent paper of McKay and Wanless [14].

## 2. On reconstructing reducible $n$-quasigroups

In what follows the constant tuples $\bar{o}, \bar{\theta}$ may be considered as all-zero tuples. From this point of view, the main result of this section states that a reducible $n$-quasigroup is uniquely specified by its values on the 'shell', where the 'shell' is the set of $n$-ples with at least one zero. Lemma 1 and its corollary concern the case when the groups of variables in the decomposition of a reducible $n$-quasigroup are fixed. In Theorem 1 the groups of variables are not specified; we have to require $n \geqslant 4$ in this case.
Lemma 1 (a representation of a reducible $n$-quasigroup by the superposition of retracts). Let $h$ and $g$ be an $(n-m+1)$ - and m-quasigroups, let $\bar{o} \in \Sigma^{m-1}, \bar{\theta} \in \Sigma^{n-m}$, and let

$$
\begin{gather*}
f(x, \bar{y}, \bar{z}) \stackrel{\text { def }}{=} h(g(x, \bar{y}), \bar{z}), \\
h_{0}(x, \bar{z}) \stackrel{\text { def }}{=} f(x, \bar{o}, \bar{z}), \quad g_{0}(x, \bar{y}) \stackrel{\text { def }}{=} f(x, \bar{y}, \bar{\theta}), \quad \delta(x) \stackrel{\text { def }}{=} f(x, \bar{o}, \bar{\theta}) \tag{1}
\end{gather*}
$$

where $x \in \Sigma, \bar{y} \in \Sigma^{m-1}, \bar{z} \in \Sigma^{n-m}$. Then

$$
\begin{equation*}
f(x, \bar{y}, \bar{z}) \equiv h_{0}\left(\delta^{-1}\left(g_{0}(x, \bar{y})\right), \bar{z}\right) \tag{2}
\end{equation*}
$$

Proof. It follows from (1) that
$h_{0}(\cdot, \bar{z}) \equiv h(g(\cdot, \bar{o}), \bar{z}), \quad g_{0}(x, \bar{y}) \equiv h(g(x, \bar{y}), \bar{\theta}), \quad \delta^{-1}(\cdot) \equiv g^{-1}\left(h^{-1}(\cdot, \bar{\theta}), \bar{o}\right)$.
Substituting these representations of $h_{0}, g_{0}, \delta^{-1}$ to (2), we can readily verify its validity.
Corollary 1. Let $q_{i n}, q_{\text {out }}, f_{\text {in }}, f_{\text {out }}: \Sigma^{2} \rightarrow \Sigma$ be some quasigroups, $q \stackrel{\text { def }}{=} q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}\right)\right), f \stackrel{\text { def }}{=} f_{\text {out }}\left(x_{1}, f_{\text {in }}\left(x_{2}, x_{3}\right)\right)$, and $\left(o_{1}, o_{2}, o_{3}\right) \in \Sigma^{3}$. Assume that for all $\left(x_{1}, x_{2}, x_{3}\right) \in \Sigma^{3}$ it holds

$$
q\left(o_{1}, x_{2}, x_{3}\right)=f\left(o_{1}, x_{2}, x_{3}\right), \quad q\left(x_{1}, o_{2}, x_{3}\right)=f\left(x_{1}, o_{2}, x_{3}\right) .
$$

Then $q(\bar{x})=f(\bar{x})$ for all $\bar{x} \in \Sigma^{3}$.
Theorem 1. Let $q, f: \Sigma^{n} \rightarrow \Sigma$ be reducible $n$-quasigroups, where $n \geqslant 4$; and let $o_{1}^{n} \in \Sigma^{n}$. Assume that for all $i \in\{1, \ldots, n\}$ and for all $x_{1}^{n} \in \Sigma^{n}$ it holds

$$
\begin{equation*}
q\left(x_{1}^{i-1}, o_{i}, x_{i+1}^{n}\right)=f\left(x_{1}^{i-1}, o_{i}, x_{i+1}^{n}\right) \tag{3}
\end{equation*}
$$

Then $q\left(x_{1}^{n}\right)=f\left(x_{1}^{n}\right)$ for all $x_{1}^{n} \in \Sigma^{n}$.

Proof. (*) We first proof the claim for $n=4$. Without loss of generality (up to coordinate permutation and/or interchanging $q$ and $f$ ), we can assume that one of the following holds for some quasigroups $q_{\text {in }}, q_{\text {out }}, f_{\text {in }}, f_{\text {out }}$ :

Case 1) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}, x_{4}\right)\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(x_{1}, f_{\text {in }}\left(x_{2}, x_{3}, x_{4}\right)\right)$;
Case 2) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}, x_{4}\right)\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(x_{1}, f_{\text {in }}\left(x_{2}, x_{3}\right), x_{4}\right)$;
Case 3) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}\right), x_{4}\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(x_{1}, f_{\text {in }}\left(x_{2}, x_{3}\right), x_{4}\right)$;
Case 4) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}, x_{4}\right)\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(f_{\text {in }}\left(x_{1}, x_{2}, x_{3}\right), x_{4}\right)$;
Case 5) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}, x_{4}\right)\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(f_{\text {in }}\left(x_{1}, x_{4}\right), x_{2}, x_{3}\right)$;
Case 6) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, x_{2}, q_{\text {in }}\left(x_{3}, x_{4}\right)\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(x_{1}, f_{\text {in }}\left(x_{2}, x_{3}\right), x_{4}\right)$;
Case 7) $q\left(x_{1}^{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}\right), x_{4}\right), f\left(x_{1}^{4}\right)=f_{\text {out }}\left(f_{\text {in }}\left(x_{1}, x_{4}\right), x_{2}, x_{3}\right)$.
$1,2,3)$ Take an arbitrary $x_{4}$ and denote $q^{\prime}\left(x_{1}, x_{2}, x_{3}\right) \stackrel{\text { def }}{=} q\left(x_{1}, x_{2}, x_{3}, x_{4}\right)$ and $f^{\prime}\left(x_{1}, x_{2}, x_{3}\right) \stackrel{\text { def }}{=} f\left(x_{1}, x_{2}, x_{3}, x_{4}\right)$. Then, by Corollary 1 , we have $q^{\prime}(\bar{x})=$ $f^{\prime}(\bar{x})$ for all $\bar{x} \in \Sigma^{3}$; this proves the statement.
4) Fixing $x_{4}:=o_{4}$ and applying (3) with $i=4$, we have

$$
f_{\text {out }}\left(f_{\text {in }}\left(x_{1}, x_{2}, x_{3}\right), o_{4}\right)=q_{\text {out }}\left(x_{1}, q_{\text {in }}\left(x_{2}, x_{3}, o_{4}\right)\right)
$$

which leads to the representation $f_{\text {in }}\left(x_{1}, x_{2}, x_{3}\right)=h_{\text {out }}\left(x_{1}, h_{\text {in }}\left(x_{2}, x_{3}\right)\right)$ where $h_{\text {out }}\left(x_{1}, \cdot\right) \stackrel{\text { def }}{=} f_{\text {out }}^{-1}\left(q_{\text {out }}\left(x_{1}, \cdot\right), o_{4}\right)$ and $h_{\text {in }}\left(x_{2}, x_{3}\right) \stackrel{\text { def }}{=} q_{\text {in }}\left(x_{2}, x_{3}, o_{4}\right)$. Using this representation, we find that $f$ satisfies the condition of Case 2) for some $f_{\text {in }}, f_{\text {out }}$. So, the situation is reduced to the already-considered case.
5) Fixing $x_{4}:=o_{4}$ and using (3), we obtain the decomposition $f_{\text {out }}(\cdot, \cdot, \cdot)=$ $h_{\text {out }}\left(\cdot, h_{\text {in }}(\cdot, \cdot)\right)$ for some $h_{\text {in }}, h_{\text {out }}$. We find that $q$ and $f$ satisfy the conditions of Case 2).
6) Fixing $x_{4}:=o_{4}$ and using (3), we get the decomposition $q_{o u t}(\cdot, \cdot, \cdot)=$ $h_{\text {out }}\left(\cdot, h_{\text {in }}(\cdot, \cdot)\right)$. Then, we again reduce to Case 2).
7) Fixing $x_{4}:=o_{4}$ we derive the decomposition $f_{\text {out }}(\cdot, \cdot, \cdot)=h_{\text {out }}\left(\cdot, h_{\text {in }}(\cdot, \cdot)\right)$, which leads to Case 3 ).
$\left.{ }^{* *}\right)$ Assume $n>4$. It is straightforward to show that we always can choose four indexes $1 \leqslant i<j<k<l \leqslant n$ such that for all $x_{1}^{i-1}, x_{i+1}^{j-1}$, $x_{j+1}^{k-1}, x_{k+1}^{l-1}, x_{l+1}^{n}$ the 4-quasigroups

$$
\begin{aligned}
& q_{x_{1}^{i-1} x_{i+1}^{j-1} x_{j+1}^{k-1} x_{k+1}^{l-1} x_{l+1}^{n}}\left(x_{i}, x_{j}, x_{k}, x_{l}\right) \stackrel{\text { def }}{=} q\left(x_{1}^{n}\right) \\
& f_{x_{1}^{i-1} x_{i+1}^{j-1} x_{j+1}^{k-1} x_{k+1}^{l-1} x_{l+1}^{n}}^{\prime}\left(x_{i}, x_{j}, x_{k}, x_{l}\right) \stackrel{\text { def }}{=} f\left(x_{1}^{n}\right)
\end{aligned}
$$

are reducible. Since these 4-quasigroups satisfy the hypothesis of the lemma, they are identical, according to $\left(^{*}\right)$. Since they coincide for every values of the parameters, we see that $q$ and $f$ are also identical.

Remark 1. If $n=3$ then the claim of Lemma 1 can fail. For example, the reducible 3-quasigroups $q\left(x_{1}^{3}\right) \stackrel{\text { def }}{=}\left(x_{1} * x_{2}\right) * x_{3}$ and $f\left(x_{1}^{3}\right) \stackrel{\text { def }}{=} x_{1} *\left(x_{2} * x_{3}\right)$ where $*$ is a binary quasigroup with an identity element 0 (i.e., a loop) coincide if $x_{1}=0, x_{2}=0$, or $x_{3}=0$; but they are not identical if $*$ is nonassociative.

## 3. Subquasigroup

Let $q: \Sigma^{n} \rightarrow \Sigma$ be an $n$-quasirgoup and $\Omega \subset \Sigma$. If $g=\left.q\right|_{\Omega^{n}}$ is an $n$ quasirgoup then we will say that $g$ is a subquasigroup of $q$ and $q$ is $\Omega$-closed.
Lemma 2. For each finite $\Sigma$ with $|\Sigma|=k$ and $\Omega \subset \Sigma$ with $|\Omega| \leqslant\lfloor k / 2\rfloor$ there exists a reducible $n$-quasigroup $q: \Sigma^{n} \rightarrow \Sigma$ with a subquasigroup $g: \Omega^{n} \rightarrow \Omega$.

Proof. By Ryser theorem on completion of a Latin $s \times r$ rectangular up to a Latin $k \times k$ square (2-quasigroup) [16], there exists a $\Omega$-closed 2 -quasigroup $q: \Sigma^{2} \rightarrow \Sigma$.

To be constructive, we suggest a direct formula for the case $\Sigma=\{0, \ldots$, $k-1\}, \Omega=\{0, \ldots, r-1\}$ where $k \geqslant 2 r$ and $k-r$ is odd:

$$
\begin{array}{rlrl}
q_{k, r}(i, j) & =(i+j) \bmod r, & & i<r, j<r ; \\
q_{k, r}(r+i, j) & =(i+j) \bmod (k-r)+r, & & j<r ; \\
q_{k, r}(i, r+j) & =(2 i+j) \bmod (k-r)+r, & & i<r ; \\
q_{k, r}(r+i, r+j) & = \begin{cases}(i-j) \bmod (k-r) & \\
\text { if }(i-j) \bmod (k-r)<r, \\
(2 i-j) \bmod (k-r)+r & \\
\text { otherwise. } .\end{cases}
\end{array}
$$

In the following four examples the second and the fourth value arrays correspond to $q_{5,2}$ and $q_{7,2}$ :

4: | 0 | 1 | 2 | 3 |
| :--- | :--- | :--- | :--- |
| 1 | 0 | 3 | 2 |
| 2 | 3 | 0 | 1 |
| 3 | 2 | 1 | 0 |



$6: |$| 0 | 1 | 2 | 3 | 4 | 5 |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| 1 | 0 | 3 | 2 | 5 | 4 |
| 4 | 5 | 0 | 1 | 2 | 3 |
| 5 | 4 | 1 | 0 | 3 | 2 |
| 2 | 3 | 4 | 5 | 0 | 1 |
| 3 | 2 | 5 | 4 | 1 | 0 |


| 0 | 1 | 2 | 3 | 4 | 5 | 6 |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| 1 | 0 | 3 | 4 | 5 | 6 | 2 |
| 2 | 4 | 0 | 1 | 6 | 3 | 5 |
| 3 | 5 | 6 | 0 | 1 | 2 | 4 |
| 4 | 6 | 5 | 2 | 0 | 1 | 3 |
| 5 | 2 | 4 | 6 | 3 | 0 | 1 |
| 6 | 3 | 1 | 5 | 2 | 4 | 0 |

Now, the statement follows from the obvious fact that a superposition of $\Omega$-closed 2 -quasigroups is an $\Omega$-closed $n$-quasigroup.

The next obvious lemma is a suitable tool for obtaining a large number of $n$-quasigroups, most of which are irreducible.

Lemma 3 (switching subquasigroups). Let $q: \Sigma^{n} \rightarrow \Sigma$ be an $\Omega$-closed $n$-quasigroup with a subquasigroup $g: \Omega^{n} \rightarrow \Omega, g=\left.q\right|_{\Omega^{n}}, \Omega \subset \Sigma$. And let $h: \Omega^{n} \rightarrow \Omega$ be another $n$-quasigroup of order $|\Omega|$. Then

$$
f(\bar{x}) \stackrel{\text { def }}{=} \begin{cases}h(\bar{x}) & \text { if } \bar{x} \in \Omega^{n}  \tag{5}\\ q(\bar{x}) & \text { if } \bar{x} \notin \Omega^{n}\end{cases}
$$

is an n-quasigroup of order $|\Sigma|$.

## 4. Irreducible $n$-quasigroups

Lemma 4. A subquasigroup of a reducible n-quasigroup is reducible.
Proof. Let $f: \Sigma^{n} \rightarrow \Sigma$ be a reducible $\Omega$-closed $n$-quasigroup. Without loss of generality we assume that

$$
f(x, \bar{y}, \bar{z}) \equiv h(g(x, \bar{y}), \bar{z})
$$

for some ( $n-m+1$ )- and $m$-quasigroups $h$ and $g$ where $1<m<n$. Take $\bar{o} \in \Omega^{m-1}$ and $\theta \in \Omega^{n-m}$. Then the quasigroups $h_{0}, g_{0}$, and $\delta$ defined by (1) are $\Omega$-closed. Therefore, the representation (2) proves that $\left.f\right|_{\Omega^{n}}$ is reducible.
Theorem 2. For each $n \geqslant 3$ and $k \geqslant 4$ there exists an irreducible $n$-quasigroup of order $k$.
Proof. (*) First we consider the case $n \geqslant 4$. By Lemma 2 we can construct a reducible $n$-quasigroup $q:\{0, \ldots, k-1\}^{n} \rightarrow\{0, \ldots, k-1\}$ of order $k$ with a subquasigroup $g:\{0,1\}^{n} \rightarrow\{0,1\}$ of order 2 . Let $h:\{0,1\}^{n} \rightarrow\{0,1\}$ be the $n$-quasigroup of order 2 different from $g$; and let $f$ be defined by (5). By Theorem 1 with $\bar{o}=(2, \ldots, 2)$, the $n$-quasigroup $f$ is irreducible.
$\left({ }^{* *}\right) n=3, k=4,5,6,7$. In each of these cases we will construct an irreducible 3 -quasigroup $f$, omitting the verification, which can be done, for example, using the formulas (1), (2). Let quasigroups $q_{4,2}, q_{5,2}, q_{6,2}$, and $q_{7,2}$ be defined by the value arrays (4). For each case $k=4,5,6,7$ we define the ternary quasigroup $q\left(x_{1}, x_{2}, x_{3}\right) \stackrel{\text { def }}{=} q_{k, 2}\left(q_{k, 2}\left(x_{1}, x_{2}\right), x_{3}\right)$, which have the subquasigroup $\left.q\right|_{\{0,1\}^{3}}\left(x_{1}, x_{2}, x_{3}\right)=x_{1}+x_{2}+x_{3} \bmod 2$. Using (5), we replace this subquasigroup by the ternary quasigroup $h\left(x_{1}, x_{2}, x_{3}\right)=$ $x_{1}+x_{2}+x_{3}+1 \bmod 2$. The resulting ternary quasigroup $f$ is irreducible.
$\left(^{* * *}\right) n=3,8 \leqslant k<\infty$. Using Lemma 2 , Lemma 3 , and $\left({ }^{* *}\right)$, we can easily construct a ternary quasigroup of order $k \geqslant 8$ with an irreducible subquasigroup of order 4. By Lemma 4, such quasigroup is irreducible.
(****) The case of infinite order. Let $q: \Sigma_{\infty}^{n} \rightarrow \Sigma_{\infty}$ be an $n$-quasigroup of infinite order $K$ and $g: \Sigma^{n} \rightarrow \Sigma$ be any irreducible $n$-quasigroup of finite order (say, 4). Then, by Lemma 4, their direct product $g \times q:\left(\Sigma \times \Sigma_{\infty}\right)^{n} \rightarrow\left(\Sigma \times \Sigma_{\infty}\right)$ defined as

$$
g \times q\left(\left[x_{1}, y_{1}\right], \ldots,\left[x_{n}, y_{n}\right]\right) \stackrel{\text { def }}{=}\left[g\left(x_{1}, \ldots, x_{n}\right), q\left(y_{1}, \ldots, y_{n}\right)\right]
$$

is an irreducible $n$-quasigroup of order $K$.
Remark 2. Using the same arguments, it is easy to construct for any $n \geqslant 4$ and $k \geqslant 4$ an irreducible $n$-quasigroup of order $k$ such that fixing one argument (say, the first) by (say) 0 leads to an ( $n-1$ )-quasigroup that is also irreducible. This simple observation naturally blends with the following context. Let $\kappa(q)$ be the maximal number such that there is an irreducible $\kappa(q)$-quasigroup that can be obtained from $q$ or one of its inverses by fixing $n-\kappa(q)>0$ arguments. In this remark we observe that (for any $n$ and $k$ when the question is nontrivial) there is an irreducible $n$-quasigroup $q$ with
$\kappa(q)=n-1$. In [10] for $k: 4$ and even $n \geqslant 4$ an irreducible $n$-quasigroup with $\kappa(q)=n-2$ is constructed. In [9, 12] it is shown that $\kappa(q) \leqslant n-3$ (if $k$ is prime then $\kappa(q) \leqslant n-2$ ) implies that $q$ is reducible.

## 5. On the number of $n$-quasigroups, I

We first consider a simple bound on the number of $n$-quasigroups of composite order.
Proposition 1. The number $|Q(n, s r)|$ of $n$-quasigroups of composite order sr satisfies

$$
\begin{equation*}
|Q(n, s r)| \geqslant|Q(n, r)| \cdot|Q(n, s)|^{r^{n}}>|Q(n, s)|^{r^{n}} \tag{6}
\end{equation*}
$$

Proof. Let $g: Z_{r}^{n} \rightarrow Z_{r}$ be an arbitrary $n$-quasigroup of order $r$; and let $\omega\langle\cdot\rangle$ be an arbitrary function from $Z_{r}^{n}$ to the set $Q(n, s)$ of all $n$-quasigroups of order $s$. It is straightforward that the following function is an $n$-quasigroup of order $s r$ :

$$
f\left(z_{1}^{n}\right) \stackrel{\text { def }}{=} g\left(y_{1}^{n}\right) \cdot s+\omega\left\langle y_{1}^{n}\right\rangle\left(x_{1}^{n}\right) \quad \text { where } y_{i} \stackrel{\text { def }}{=}\left\lfloor z_{i} / s\right\rfloor, \quad x_{i} \stackrel{\text { def }}{=} z_{i} \bmod s
$$

Moreover, different choices of $\omega\langle\cdot\rangle$ result in different $n$-quasigroups. So, this construction, which is known as the $\omega$-product of $g$, obviously provides the bound (6).

If the order is divided by 2 or 3 then the bound (6) is the best known. Substituting the known values $|Q(n, 2)|=2$ and $|Q(n, 3)|=3 \cdot 2^{n}$, we get

Corollary 2. If $k: 2$ then $|Q(n, k)| \geqslant 2^{(k / 2)^{n}}$;

$$
\text { if } k: 3 \text { then }|Q(n, k)| \geqslant\left(3 \cdot 2^{n}\right)^{(k / 3)^{n}}>2^{n(k / 3)^{n}}
$$

The next statement is weaker than the bound considered in the next section. Nevertheless, it provides simplest arguments showing that the number of $n$-quasigroup of fixed order $k$ grows double-exponentially, even for prime $k \geqslant 8$. The cases $k=5$ and $k=7$ will be covered in the next section.
Proposition 2. The number $|Q(n, k)|$ of $n$-quasigroups of order $k \geqslant 8$ satisfies

$$
\begin{equation*}
|Q(n, k)| \geqslant 2^{\lfloor k / 4\rfloor^{n}} \tag{7}
\end{equation*}
$$

Proof. By Lemma 2, there is an $n$-quasigroup of order $k$ with subquasigroup of order $2\lfloor k / 4\rfloor$. This subquasigroup can be switched (see Lemma 3) in $|Q(n, 2\lfloor k / 4\rfloor)|$ ways. By Proposition 1, we have

$$
|Q(n, 2\lfloor k / 4\rfloor)| \geqslant|Q(n, 2)|^{\lfloor k / 4\rfloor^{n}}=2^{\lfloor k / 4\rfloor^{n}}
$$

Clearly, these calculations have sense only if $\lfloor k / 4\rfloor>1$, i. e., $k \geqslant 8$.

## 6. On the number of $n$-quasigroups, II

In this section we continue using the same general switching principle as in previous ones: independent changing the values of $n$-quasigroups on disjoint subsets of $\Sigma^{n}$. We improve the lower bound in the cases when the order is not divided by 2 or 3 ; in particular, we establish a double-exponential lower bound on the number of $n$-quasigroups of orders 5 and 7 .

We say that a nonempty set $\Theta \subset \Sigma^{n}$ is an ab-component or a switching component of an $n$-quasigroup $q$ iff
(a) $q(\Theta)=\{a, b\}$ and
(b) the function $q \Theta: \Sigma^{n} \rightarrow \Sigma$ defined as follows is an $n$-quasigroup too:

$$
q \Theta(\bar{x}) \stackrel{\text { def }}{=} \begin{cases}q(\bar{x}) & \text { if } \bar{x} \notin \Theta \\ b & \text { if } \bar{x} \in \Theta \text { and } q(\bar{x})=a \\ a & \text { if } \bar{x} \in \Theta \text { and } q(\bar{x})=b\end{cases}
$$

For example, $\{(0,0),(0,1),(1,0),(1,1)\}$ and $\{(2,2),(2,3),(3,3),(3,4)$, $(4,2),(4,4)\}$ are 01-components in (4.5).
Remark 3. From some point of view, it is naturally to require also $\Theta$ to be inclusion-minimal, i.e., (c) $\Theta$ does not have a nonempty proper subset that satisfies (a) and (b). Although in what follows all ab-components satisfy (c), formally we do not use it.

Lemma 5. Let an n-quasigroup $q$ have $s$ pairwise disjoint switching components $\Theta_{1}, \ldots, \Theta_{s}$ (note that we do not require them to be ab-components for common $a, b)$. Then $|Q(n,|\Sigma|)| \geqslant 2^{s}$.
Proof. Indeed, denoting $q \Theta^{0} \stackrel{\text { def }}{=} q$ and $q \Theta^{1} \stackrel{\text { def }}{=} q \Theta$, we have $2^{s}$ distinct $n$ quasigroups $q \Theta_{1}^{t_{1}} \ldots \Theta_{s}^{t_{s}},\left(t_{1}, \ldots, t_{s}\right) \in\{0,1\}^{s}$.

### 6.1. The order 5

In this section, we consider the $n$-quasigroups of order 5 , the only case, when the other our bounds do not guarantee the double-exponential growth of the number of $n$-quasigroups as $n \rightarrow \infty$. Of course, the way that we use for the order 5 works for any other order $k>3$, but the bound obtained is worse than (6) provided $k$ is composite, worse than (7) provided $k \geqslant 8$, and worse than (8) provided $k \geqslant 6$. The bound is based on the following straightforward fact:
Lemma 6. Let $\{0,1\}^{n}$ be a 01-component of an $n$-quasigroup $q$. For every $i \in\{1, \ldots, n\}$ let $q_{i}$ be an $n_{i}$-quasigroup and let $\Theta_{i}$ be its 01-component. Then $\Theta_{1} \times \ldots \times \Theta_{n}$ is a 01-component of the $\left(n_{1}+\ldots+n_{n}\right)$-quasigroup
$f\left(x_{1,1}, \ldots, x_{1, n_{1}}, x_{2,1}, \ldots, x_{n, n_{n}}\right) \stackrel{\text { def }}{=} q\left(q_{1}\left(x_{1,1}, \ldots, x_{1, n_{1}}\right), \ldots, q_{n}\left(x_{n, 1}, \ldots, x_{n, n_{n}}\right)\right)$.

For a quasigroup $q: \Sigma^{2} \rightarrow \Sigma$ denote $q^{1} \stackrel{\text { def }}{=} q, q^{2}\left(x_{1}, x_{2}, x_{3}\right) \stackrel{\text { def }}{=} q\left(x_{1}, q^{1}\left(x_{2}, x_{3}\right)\right)$, $\ldots, q^{i}\left(x_{1}, x_{2}, \ldots, x_{i+1}\right) \stackrel{\text { def }}{=} q\left(x_{1}, q^{i-1}\left(x_{2}, \ldots, x_{i+1}\right)\right)$.
Proposition 3. If $n=3 m$ then $|Q(n, 5)| \geqslant 2^{3^{m}}$; if $n=3 m+1$ then $|Q(n, 5)| \geqslant 2^{4 \cdot 3^{m-1}} ;$ if $n=3 m+2$ then $|Q(n, 5)| \geqslant 2^{2 \cdot 3^{m}}$. Roughly, for any $n$ we have

$$
|Q(n, 5)|>2^{3^{n / 3-0.072}}>e^{e^{\frac{\ln 3}{3} n-0.44}}
$$

Proof. Let $q$ be the quasigroup of order 5 with value table (4.5). Then
$\left(^{*}\right) q$ has two disjoint 01-components $D_{0} \stackrel{\text { def }}{=}\{(0,0),(0,1),(1,0),(1,1)\}$ and $D_{1} \stackrel{\text { def }}{=}\{(2,2),(2,3),(3,3),(3,4),(4,2),(4,4)\}$;
${ }^{(* *)} q^{2}$ has three mutually disjoint 01-components $T_{0} \stackrel{\text { def }}{=}\{0,1\} \times D_{0}$, $T_{1} \stackrel{\text { def }}{=}\{0,1\} \times D_{1}$, and $T_{2} \stackrel{\text { def }}{=}\left\{\left(x_{1}, x_{2}, x_{3}\right) \mid q^{2}\left(x_{1}, x_{2}, x_{3}\right) \in\{0,1\}\right\} \backslash\left(T_{0} \cup T_{1}\right) ;$
(***) $\{0,1\}^{m+1}$ is a 01 -component of $q^{m}$.
By Lemma 6,
i. the $3 m$-quasigroup defined as the superposition

$$
q^{m-1}\left(q^{2}(\cdot, \cdot, \cdot), \ldots, q^{2}(\cdot, \cdot, \cdot)\right)
$$

has $3^{m}$ components $T_{t_{1}} \times \ldots \times T_{t_{m}},\left(t_{1}, \ldots, t_{m}\right) \in\{0,1,2\}^{m}$;
ii. the $3 m+1$-quasigroup defined as the superposition

$$
q^{m}\left(q^{2}(\cdot, \cdot, \cdot), \ldots, q^{2}(\cdot, \cdot, \cdot), q(\cdot, \cdot), q(\cdot, \cdot)\right)
$$

has $3^{m-1} 4$ components $T_{t_{1}} \times \ldots \times T_{t_{m-1}} \times D_{t_{m}} \times D_{t_{m+1}},\left(t_{1}, \ldots, t_{m+1}\right) \in$ $\{0,1,2\}^{m-1} \times\{0,1\}^{2} ;$
iii. the $3 m+2$-quasigroup defined as the superposition

$$
q^{m}\left(q^{2}(\cdot, \cdot, \cdot), \ldots, q^{2}(\cdot, \cdot, \cdot), q(\cdot, \cdot)\right)
$$

has $3^{m} 2$ components $T_{t_{1}} \times \ldots \times T_{t_{m}} \times D_{t_{m+1}},\left(t_{1}, \ldots, t_{m+1}\right) \in\{0,1,2\}^{m} \times$ $\{0,1\}$.

By Lemma 5, the theorem follows.
Remark 4. If, in the proof, we consider the superposition $q^{n / 2}(q(\cdot, \cdot), \ldots$, $\left.q^{2}(\cdot, \cdot)\right)$, then we obtain the bound $|Q(n, 5)| \geqslant 2^{2^{n / 2}}$ for even $n$, which is worse because $\frac{\ln 2}{2}<\frac{\ln 3}{3}$.

### 6.2. The case of order $\geqslant 7$

In this section, we will prove the following:
Proposition 4. The number $|Q(n, k)|$ of $n$-quasigroups $\{0,1, \ldots, k-1\}^{n} \rightarrow$ $\{0,1, \ldots, k-1\}$ satisfies

$$
\begin{equation*}
|Q(n, k)| \geqslant 2^{\lfloor k / 2\rfloor\lfloor k / 3\rfloor^{n-1}}>e^{e^{\ln \lfloor k / 3\rfloor n+\ln \lfloor k / 2\rfloor-\ln \lfloor k / 3\rfloor-0.37}}>e^{e^{\ln \lfloor k / 3\rfloor n+0.038}} \tag{8}
\end{equation*}
$$

Note that this bound has no sense if $k<6$; and it is weaker than (6) if $k: 2$ or $k: 3$. The proof is based on the following straightforward fact:
Lemma 7. Let $\{c, d\} \times\{e, f\}$ be an ab-component of a quasigroup $g$. Then
(a) $\{a, b\} \times\{e, f\}$ is a cd-component of the quasigroup $g^{-}$defined by $g(x, y)=z \Leftrightarrow g^{-}(z, y)=x ;$
(b) if $\left\{a_{1}, b_{1}\right\} \times \ldots \times\left\{a_{n}, b_{n}\right\}$ is a ef-component of an n-quasigroup $q$, then $\{c, d\} \times\left\{a_{1}, b_{1}\right\} \times \ldots \times\left\{a_{n}, b_{n}\right\}$ is an ab-component of the $(n+1)$-quasigroup defined as the superposition $g(\cdot, q(\cdot, \ldots, \cdot))$.
Proof of Proposition 4. Taking into account Corollary 2, it is enough to consider only the cases of odd $k \not \equiv 0 \bmod 3$. Moreover, we can assume that $k>6$ (otherwise the statement is trivial).

Define the 2-quasigroup $q$ as

$$
\begin{aligned}
q(2 j, i) & \stackrel{\text { def }}{=} i+3 j \bmod k ; \\
q(2 j+1, i) & \stackrel{\text { def }}{=} \pi(i)+3 j \bmod k ; \\
q(2\lfloor k / 3\rfloor+j, i) & \stackrel{\text { def }}{=} \tau(i)+3 j \bmod k ; j=0, \ldots,\lfloor k / 3\rfloor-1, i=0, \ldots, k-1
\end{aligned}
$$

where $\pi, \tau$, and the remaining values of $q$ are defined by the following value table (the fourth row is used only for the case $k \equiv 2 \bmod 3$ ):

| $i$ | 0 | 1 | 2 | 3 | 4 | $\ldots$ | $k-5$ | $k-4$ | $k-3$ | $k-2$ | $k-1$ |
| ---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\pi(i)$ | 1 | 0 | 3 | 2 | 5 | $\ldots$ | $k-4$ | $k-5$ | $k-2$ | $k-1$ | $k-3$ |
| $\tau(i)$ | $k-1$ | 2 | 1 | 4 | 3 | $\ldots$ | $k-3$ | $k-4$ | 0 | $k-2$ |  |
| $q(k-2, i)$ | $k-3$ | $k-2$ | $k-1$ | 0 | 1 | $\ldots$ | $k-7$ | $k-6$ | $k-4$ | $k-5$ |  |
| $q(k-1, i)$ | $k-2$ | $k-1$ | 0 | 1 | 2 | $\ldots$ | $k-6$ | $k-5$ | $k-3$ | $k-4$ |  |

In what follows, the tables illustrate the cases $k=7$ and $k=11$.

$$
k=7: \begin{array}{|ll|ll|lll|}
\hline 0 & 1 & 2 & 3 & 4 & 5 & 6 \\
1 & 0 & 3 & 2 & 5 & 6 & 4 \\
\hline 3 & 4 & 5 & 6 & 0 & 1 & 2 \\
4 & 3 & 6 & 5 & 1 & 2 & 0 \\
\hline 6 & 2 & 1 & 4 & 3 & 0 & 5 \\
2 & 5 & 4 & 0 & 6 & 3 & 1 \\
5 & 6 & 0 & 1 & 2 & 4 & 3 \\
\hline
\end{array}
$$

|  |  | 0 1 <br> 1 0 <br>   |  | 2 3 |  | 45 |  |  | 7 |  |  | 9 |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  |  | 34 | 5 | 56 |  | 78 | 8 | 9 | 10 |  |  | 1 | 2 |
|  |  | 43 | 6 | 65 |  | 87 |  | 10 | 9 |  |  | 2 |  |
| $k=11$ : |  | 67 |  | 89 |  | 10 |  | 1 | 2 |  |  | 4 | 5 |
|  |  | 76 | 9 |  |  | 01 |  | 2 | 1 |  |  | 5 |  |

For each $j=0, \ldots,\lfloor k / 3\rfloor-1$ and $i=0, \ldots,\lfloor k / 2\rfloor-2$ the set $\{2 j, 2 j+1\} \times$ $\{2 i, 2 i+1\}$ is a $(2 i+3 j \bmod k)(2 i+3 j+1 \bmod k)$-component of such $q$. By Lemma $7(\mathrm{a})$, for the same pairs $i, j$ the set $\{2 i+3 j \bmod k, 2 i+3 j+$ $1 \bmod k\} \times\{2 i, 2 i+1\}$ is a $(2 j)(2 j+1)$-component of $g \stackrel{\text { def }}{=} q^{-}$; moreover, we can observe that for each $j$ there is one more "non-square" $(2 j)(2 j+1)$ component of $g$ which is disjoint with all considered "square" components, see the following examples (we omit the analytic description; indeed, we can
ignore this component if we do not care about the constant in the bound $\left.e^{e^{\ln \lfloor k / 3\rfloor n+c o n s t}}\right)$.

$$
k=7: \begin{array}{|ll|ll|llll}
\hline 0 & 1 & 6 & 5 & 2 & 4 & 3 \\
1 & 0 & 4 & 6 & 3 & 2 & 5 \\
\hline 5 & 4 & 0 & 1 & 6 & 3 & 2 \\
\hline 2 & 3 & 1 & 0 & 4 & 5 & 6 \\
3 & 2 & 5 & 4 & 0 & 6 & 1 \\
\hline 6 & 5 & 2 & 3 & 1 & 0 & 4 \\
4 & 6 & 3 & 2 & 5 & 1 & 0 \\
\hline
\end{array}
$$



By induction, using Lemma 7 (b), we derive that for every $j_{1}, \ldots, j_{n-1} \in$ $\{0, \ldots,\lfloor k / 3\rfloor-1\}$ and $i \in\{0, \ldots,\lfloor k / 2\rfloor-2\}$ the set

$$
\begin{aligned}
& \left\{\quad 2 j_{2}+3 j_{1} \bmod k, \quad 2 j_{2}+3 j_{1}+1 \bmod k\right\} \times \\
& \cdots \\
& \left\{2 j_{n-1}+3 j_{n-2} \bmod k, 2 j_{n-1}+3 j_{n-2}+1 \bmod k\right\} \times \\
& \left\{\begin{array}{rr}
2 i+3 j_{n-1} \bmod k, & \left.2 i+3 j_{n-1}+1 \bmod k\right\}
\end{array}\right) \times\{2 i, 2 i+1\}
\end{aligned}
$$

is a $\left(2 j_{1}\right)\left(2 j_{1}+1\right)$-component of the $n$-quasigroup $g^{n-1}$. Also, for every such $j_{1}, \ldots, j_{n-1}$ there is one more $\left(2 j_{1}\right)\left(2 j_{1}+1\right)$-component of $g^{n-1}$, which is generated by the "non-square" $\left(2 j_{n-1}\right)\left(2 j_{n-1}+1\right)$-component of $g$. In summary, $g^{n-1}$ has at least $\lfloor k / 3\rfloor^{n-1}\lfloor k / 2\rfloor$ pairwise disjoint switching components. By Lemma 5, the theorem is proved.

Summarizing Corollary 2, Propositions 3 and 4, we get the following theorem.
Theorem 3. Let a finite set $\Sigma$ of size $k>3$ be fixed. The number $|Q(n, k)|$ of $n$-quasigroups $\Sigma^{n} \rightarrow \Sigma$ satisfies the following:
(a) If $k$ is even, then $|Q(n, k)| \geqslant 2^{(k / 2)^{n}}$.
(b) If $k$ is divided by 3 , then $|Q(n, k)| \geqslant 2^{n(k / 3)^{n}}$.
(c) If $k=5$, then $|Q(n, k)| \geqslant 2^{3^{n / 3-c}}$ where $c<0.072$ depends on $n \bmod 3$.
(d) In all other cases, $|Q(n, k)| \geqslant 2^{1.5\lfloor k / 3\rfloor^{n}}$.

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# Left almost semigroups defined by a free algebra 

Qaiser Mushtaq and Muhammad Inam


#### Abstract

We have constructed LA-semigroups through a free algebra, and the structural properties of such LA-semigroups have been investegated. Moreover, the isomorphism theorems for LA-groups constructed through free algebra have been proved.


## 1. Introduction

A left almost semigroup, abbreviated as an LA-semigroup, is an algebraic structure midway between a groupoid and a commutative semigroup. The structure was introduced by M. A. Kazim and M. Naseeruddin [3] in 1972. This structure is also known as Abel-Grassmann's groupoid, abbreviated as an AG-groupoid [6] and as an invertive groupoid [1].

A groupoid $G$ with left invertive law, that is: $(a b) c=(c b) a, \forall a, b, c \in G$, is called an LA-semigroup.

An LA-semigroup satisfies the medial law: $(a b)(c d)=(a c)(b d)$. An LA-semigroup with left identity is called an $L A$-monoid.

An LA-semigroup in which either $(a b) c=b(c a)$ or $(a b) c=b(a c)$ holds for all $a, b, c, d \in G$, is called an $A G^{*}$-groupoid [6].

Let $G$ be an LA-semigroup and $a \in G$. A mapping $L_{a}: G \longrightarrow G$, defined by $L_{a}(x)=a x$, is called the left translation by $a$. Similarly, a mapping $R_{a}: G \longrightarrow G$, defined by $R_{a}(x)=x a$, is called the right translation by $a$. An LA-semigroup $G$ is called left (right) cancellative if all the left (right) translations are injective. An LA-semigroup $G$ is called cancellative if all translations are injective.

Let $X$ be a non-empty set and $W_{X}^{\prime}$ denote the free algebra over $X$ in the variety of algebras of the type $\{0, \alpha,+\}$, consisting of nullary, unary and
binary operations determined by the following identities:

$$
\begin{gathered}
(x+y)+z=x+(y+z), \quad x+y=y+x, \quad x+0=x \\
\alpha(x+y)=\alpha x+\alpha y, \quad \alpha 0=0
\end{gathered}
$$

Every element $u \in W_{X}^{\prime}$ has the form $u=\sum_{i=1}^{r} \alpha^{n_{i}} x_{i}$, where $r \geqslant 0$, and $n_{i}$ are non-negative integers. This expression is unique up to the order of the summands. Moreover $r=0$ if and only if $u=0$.

Let us define a multiplication on $W_{X}^{\prime}$ by $u \circ v=\alpha u+\alpha^{2} v$. Then the set $W_{X}^{\prime}$ is an LA-semigroup under this binary operation. We denote it by $W_{X}$. It is easy to see that $W_{X}$ is cancellative.

If $n$ is the smallest non-negative integer such that $\alpha^{n} x=x$, then $n$ is called the order of $\alpha$. The following examples show the existence of such LA-semigroups.
Example 1. Consider a field $F_{5}=\{0,1,2,3,4\}$ and define $\alpha(x)=3 x$ for all $x \in F_{5}$. Then $F_{5}$ becomes an LA-semigroup under the binary operation defined by $u \circ v=\alpha u+\alpha^{2} v, \forall u, v \in F_{5}$.

| $\circ$ | 0 | 1 | 2 | 3 | 4 |
| :--- | :--- | :--- | :--- | :--- | :--- |
| 0 | 0 | 4 | 3 | 2 | 1 |
| 1 | 3 | 2 | 1 | 0 | 4 |
| 2 | 1 | 0 | 4 | 3 | 2 |
| 3 | 4 | 3 | 2 | 1 | 0 |
| 4 | 2 | 1 | 0 | 4 | 3 |

Example 2. Let $X=\{x, y\}$ and $\alpha$ be defined as $\alpha(a)=2 a$, for all $a \in X$ and $2 \in F_{3}$. Then the following table illustrates an LA-semigroup $W_{X}$.

| $\circ$ | 0 | $x$ | $2 x$ | $y$ | $2 y$ | $x+y$ | $2 x+y$ | $x+2 y$ | $2 x+2 y$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 0 | $x$ | $2 x$ | $y$ | $2 y$ | $x+y$ | $2 x+y$ | $x+2 y$ | $2 x+2 y$ |
| $x$ | $2 x$ | 0 | $x$ | $2 x+y$ | $2 x+2 y$ | $y$ | $x+y$ | $2 y$ | $x+2 y$ |
| $2 x$ | $x$ | $2 x$ | 0 | $x+y$ | $x+2 y$ | $2 x+y$ | $y$ | $x+2 y$ | $2 y$ |
| $y$ | $2 y$ | $x+2 y$ | $2 x+2 y$ | 0 | $y$ | $x$ | $2 x$ | $x+y$ | $2 x+y$ |
| $2 y$ | $y$ | $x+y$ | $2 x+y$ | $2 y$ | 0 | $x+2 y$ | $2 x+2 y$ | $x$ | $x+y$ |
| $x+y$ | $2 x+2 y$ | $2 y$ | $x+2 y$ | $2 x$ | $2 x+y$ | 0 | $x$ | $y$ | $x+y$ |
| $2 x+y$ | $x+2 y$ | $2 x+2 y$ | $2 y$ | $x$ | $x+y$ | $2 x$ | 0 | $2 x+2 y$ | $y$ |
| $x+2 y$ | $2 x+y$ | $y$ | $x+y$ | $2 x+2 y$ | $2 x$ | $2 y$ | $x+2 y$ | 0 | $x$ |
| $2 x+2 y$ | $x+y$ | $2 x+y$ | $y$ | $x+2 y$ | $x$ | $2 x+2 y$ | $2 y$ | $2 x+2 y$ | 0 |

An LA-semigroup is called an $L A$-band [6], if all of its elements are idempotents. An LA-band can easily be constructed from a free algebra by choosing a unary operation $\alpha$ such that $\alpha+\alpha^{2}=I d_{X}$, where $I d_{X}$ denotes the identity map on $X$.

Example 3. Define a unary operation $\alpha$ as $\alpha(x)=2 x$, where $x \in F_{5}$. Then under the binary operation o defined as above, $F_{5}$ is an LA-band.

| $\circ$ | 0 | 1 | 2 | 3 | 4 |
| :--- | :--- | :--- | :--- | :--- | :--- |
| 0 | 0 | 4 | 3 | 2 | 1 |
| 1 | 2 | 1 | 0 | 4 | 3 |
| 2 | 4 | 3 | 2 | 1 | 0 |
| 3 | 1 | 0 | 4 | 3 | 2 |
| 4 | 3 | 2 | 1 | 0 | 4 |

An LA-semigroup ( $G, \cdot$ ) is called an LA-group [5], if
(i) there exists $e \in G$ such that $e a=a$ for every $a \in G$,
(ii) for every $a \in G$ there exists $a^{\prime} \in G$ such that $a^{\prime} a=e$.

A subset $I$ of an LA-semigroup ( $G, \cdot \cdot$ ) is called a left (right) ideal of $G$, if $G I \subseteq I(I G \subseteq I)$, and $I$ is called a two sided ideal of $G$ if it is left and right ideal of $G$. An LA-semigroup is called left (right) simple, if it has no proper left (right) ideals. Consequently, an LA-semigroup is simple if it has no proper ideals.

Theorem 1. A cancellative LA-semigroup is simple.
Proof. Let $G$ be a cancellative LA-semigroup. Suppose that $G$ has a proper left ideal $I$. Then by definition $G I \subseteq I$ and so $I$ being its proper ideal, is a proper LA-subsemigroup of $G$. If $g \in G \backslash I$, then $g i \in G I$, for all $i \in I$. But $G I \subseteq I$, so there exists an $i^{\prime} \in I$, such that $g i=i^{\prime}$. Since $G$ is cancellative so is then $I$. This implies that all the right and left translations are bijective. Therefore there exists $i_{1} \in I$, such that $L_{i_{1}}(i)=i^{\prime}$. This implies that $g i=i_{1} i$. By applying the right cancellation, we obtain $g=i_{1}$. This implies that $g \in I$, which contradicts our supposition. Hence $G$ is simple.

Corollary 1. An LA-semigroup defined by a free algebra is simple.
Theorem 2. If $G$ is a right (left) cancellative LA-semigroup, then $G^{2}=G$.
Proof. Let $G$ be a right (left) cancellative LA-semigroup. Then all the right (left) translations are bijective. This implies that for each $x \in G$, there exist some $y, z \in G$ such that $R_{y}(z)=x\left(L_{y}(z)=x\right)$. Hence $G^{2}=G$.

Corollary 2. An $A G^{*}$-groupoid cannot be defined by a free algebra.

Proof. It has been proved in [6], that if $G$ is an $\mathrm{AG}^{*}$-groupoid then $G^{2}$ is a commutative semigroup. Moreover, if $G$ is a right (left) cancellative LA-semigroup, then $G^{2}=G$.

We now define a subset $T_{x}$ of $W_{X}$ such that $T_{x}=\left\{\sum_{i=1}^{r} \alpha^{n_{i}} x \mid x \in X\right\}$.
Theorem 3. $T_{x}$ is an LA-subsemigroup of $W_{X}$.
Proof. It is sufficient to show that $T_{x}$ is closed under the operation o. Let $u, v \in T_{x}$. Then $u=\sum_{i=1}^{n} \alpha^{n_{i}} x, v=\sum_{i=1}^{m} \alpha^{n_{i}} x$, and so

$$
\begin{aligned}
u \circ v & =\alpha(u)+\alpha^{2}(v)=\alpha\left(\sum_{i=1}^{n} \alpha^{n_{i}} x\right)+\alpha^{2}\left(\sum_{i=1}^{m} \alpha^{n_{i}} x\right) \\
& =\left(\sum_{i=1}^{n} \alpha^{n_{i}+1}+\sum_{i=1}^{m} \alpha^{n_{i}+2}\right) x=\sum_{i=1}^{r} \alpha^{m_{i}} x,
\end{aligned}
$$

where $r=n+m, m_{i}=n_{i}+1$ for $i \leqslant n$ and $m_{i}=n_{i}+2$ for $i>n$.
Theorem 4. If $X=\left\{x_{1}, x_{2}, \ldots, x_{n}\right\}$, then $W_{X}=T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{n}}$.
Proof. Every element $u \in W_{X}$ is of the form $u=\sum_{i=1}^{r} \alpha^{n_{i}} x_{i}$, where $r$ and $n_{i}$ are non-negative integers. This expression is unique up to the order of the summands. This implies that $W_{X}=T_{x_{1}}+T_{x_{2}}+\ldots+T_{x_{n}}$. To complete the proof it is sufficient to show that $T_{x_{i}} \cap T_{x_{j}}=\{0\}$, for $i \neq j$. Let $u \in T_{x_{i}} \cap T_{x_{j}}$, such that $u \neq 0$. Then $u \in T_{x_{i}}$ and $u \in T_{x_{j}}$. This is possible only if $x_{i}=x_{j}$. Which is a contradiction to the fact that $x_{i} \neq x_{j}$. Hence the proof.

Proposition 1. The direct sum of any $T_{x_{i}}$ and $T_{x_{j}}$ for $i \neq j$ is an $L A$ subsemigroup of $W_{X}$.

Proof. The proof is straightforward.
Theorem 5. The direct sum of any finite number of $T_{x_{i}}$ 's is an $L A$ subsemigroup of $W_{X}$.

Proof. The proof follows directly by induction.
Theorem 6. The set $W_{X} / T_{x}$ of all right (left) cosets of $T_{x}$ in $W_{X}$ is an LA-semigroup.

Proof. Let $W_{X} / T_{x}=\left\{u \circ T_{x} \mid u \in W_{X}\right\}$, and $u \circ T_{x}, v \circ T_{x} \in W_{X} / T_{x}$. Then by the medial law $\left(u \circ T_{x}\right) \circ\left(v \circ T_{x}\right)=(u \circ v) \circ T_{x} \circ T_{x}$. But $T_{x} \circ T_{x}=$ $T_{x}$. Hence $\left(u \circ T_{x}\right) \circ\left(v \circ T_{x}\right)=(u \circ v) \circ T_{x} \in W_{X} / T_{x}$.

Let $u \circ T_{x}, v \circ T_{x}, w \circ T_{x} \in W_{X} / T_{x}$. Then

$$
\begin{aligned}
\left(\left(u \circ T_{x}\right) \circ\left(v \circ T_{x}\right)\right) \circ\left(w \circ T_{x}\right) & =\left((u \circ v) \circ T_{x}\right) \circ w \circ T_{x} \\
& =((u \circ v) \circ w) \circ T_{x}=((w \circ v) \circ u) \circ T_{x} \\
& =\left(\left(w \circ T_{x}\right) \circ\left(v \circ T_{x}\right)\right) \circ\left(u \circ T_{x}\right)
\end{aligned}
$$

implies that $W_{X} / T_{x}$ is an LA-simigroup.
Remark 1. $\alpha\left(T_{x}\right)=T_{x}$.
Proposition 2. For any $T_{x} \leq W_{X}$ and $v \in W_{X}$ we have
(a) $T_{x} \circ v=(\alpha(v)) \circ T_{x}$,
(b) $T_{x} \circ\left(T_{x} \circ v\right)=\alpha^{2}\left(T_{x} \circ v\right)=\alpha^{3}\left(v \circ T_{x}\right)$,
(c) $\left(T_{x} \circ v\right) \circ T_{x}=\alpha\left(T_{x} \circ v\right)=\alpha^{2}\left(v \circ T_{x}\right)$,
(d) $T_{x} \circ v=\alpha\left(v \circ T_{x}\right)$.

Proof. The proof is straightforward.
Theorem 7. $W_{X} / T_{x_{i}}=\left\{v \circ T_{x_{i}}: v \in W_{X}\right\}$ forms a partition of $W_{X}$.
Proof. We shall show that $u \circ T_{x_{i}} \cap v \circ T_{x_{i}}=\emptyset$ for $u \neq v$, and $W_{X}=$ $\cup_{v \in W_{X}} v \circ T_{x_{i}}$. Let $w \in v \circ T_{x_{i}} \cap u \circ T_{x_{i}}$. Then $w \in v \circ T_{x_{i}}$ and $w \in u \circ T_{x_{i}}$. This implies that $w=v \circ t_{1}$ and $w=u \circ t_{2}$, where $t_{1}, t_{2} \in T_{x_{i}}$. This implies $v \circ t_{1}=u \circ t_{2}$. Hence $\alpha(v)+\alpha^{2}\left(t_{1}\right)=\alpha(u)+\alpha^{2}\left(t_{2}\right)$, which further gives $\alpha(v)=\alpha(u)+\alpha^{2}\left(t_{2}\right)-\alpha^{2}\left(t_{1}\right)$ where $\alpha^{2}\left(t_{2}\right)-\alpha^{2}\left(t_{1}\right) \in T_{x_{i}}$.

Now $\alpha(v) \in \alpha(u)+T_{x_{i}}$ yields $\alpha(v)+T_{x_{i}} \subseteq \alpha(u)+T_{x_{i}}$, i.e., $v \circ T_{x_{i}} \subseteq$ $u \circ T_{x_{i}}$. Similarly, $u \circ T_{x_{i}}=v \circ T_{x_{i}}$. Hence $v \circ T_{x_{i}} \cap u \circ T_{x_{i}}=\emptyset$. Obviously, $\cup_{v \in W_{X}} v \circ T_{x_{i}} \subseteq W_{X}$.

Conversely, let $t \in W_{X}$. Then $t=\sum_{i=1}^{r} \alpha^{n_{i}} x_{i}$ implies that

$$
\begin{aligned}
t & =\alpha^{n_{1}} x_{1}+\alpha^{n_{2}} x_{2}+\ldots+\alpha^{n_{r}} x_{r} \\
& =\alpha^{n_{i}} x_{i}+\alpha^{n_{1}} x_{1}+\alpha^{n_{2}} x_{2}+\ldots+\alpha^{n_{i-1}} x_{i-1}+\alpha^{n_{i+1}} x_{i+1}+\ldots+\alpha^{n_{r}} x_{r}
\end{aligned}
$$

If $\alpha^{n_{1}} x_{1}+\alpha^{n_{2}} x_{2}+\ldots+\alpha^{n_{i-1}} x_{i-1}+\alpha^{n_{i+1}} x_{i+1}+\ldots+\alpha^{n_{r}} x_{r}=u$, then $t=\alpha^{n_{i}} x_{i}+u, \alpha^{n_{i}} x_{i} \in T_{x_{i}}$. Now $t=\alpha^{n_{i}} x_{i}+u \in T_{x_{i}}+u=\alpha(u)+T_{x_{i}}=$ $\alpha(u)+\alpha^{2}\left(T_{x_{i}}\right)=u \circ T_{x_{i}} \in \cup_{v \in W_{X}} v \circ T_{x_{i}}$ implies $W_{X} \subseteq \cup_{v \in W_{X}} v \circ T_{x_{i}}$. Hence $W_{X}=\cup_{v \in W_{X}} v \circ T_{x}$.

Theorem 8. The order of $T_{x_{i}}$ divides the order of $W_{X}$.

Proof. If $X$ is a finite non-empty set then $W_{X}$ is also finite. This implies that the set of all the right (left) cosets of $T_{x_{i}}$ in $W_{X}$ is finite.

Let $W_{X} / T_{x_{i}}=\left\{v_{1} \circ T_{x_{i}}, v_{2} \circ T_{x_{i}}, \ldots, v_{r} \circ T_{x_{i}}\right\}$. Then by virtue of Theorem $7, W_{X}=v_{1} \circ T_{x_{i}} \cup v_{2} \circ T_{x_{i}} \cup \ldots \cup v_{r} \circ T_{x_{i}}$. This implies that $\left|W_{X}\right|=\left|v_{1} \circ T_{x_{i}}\right|+\left|v_{2} \circ T_{x_{i}}\right|+\ldots+\left|v_{r} \circ T_{x_{i}}\right|$. Thus $\left|W_{X}\right|=r\left|T_{x_{i}}\right|$. Hence $\left|W_{X}\right|=\left[T_{x_{i}}, W_{X}\right]\left|T_{x_{i}}\right|$, where $\left[T_{x_{i}}, W_{X}\right]$ denotes the number of cosets of $T_{x_{i}}$ in $W_{X}$.

Theorem 9. If $X$ is a non-empty finite set having $r$ number of elements and the order of $T_{x_{i}}$ is $n$, then $\left|W_{X}\right|=n^{r}$.

Proof. Since it has already been proved that $W_{X}=T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{r}}$ for $X=\left\{x_{1}, x_{2}, \ldots, x_{r}\right\}$, it is sufficient to show that $\left|T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{r}}\right|=$ $n^{r}$. We apply induction on $r$. Let $r=2$, that is, $W_{X}=T_{x_{1}} \oplus T_{x_{2}}$. Construct the multiplication table of $T_{x_{1}}$ and write all the elements of $T_{x_{2}}$ except 0 in the index row and in the index column. Then the number of elements in the index row or column row is $2 n-1$. We see from the multiplication table that when the elements of $T_{x_{1}}$ are multiplied by the elements of $T_{x_{2}}$ some new elements appear in the table, which are of the form $u \circ v=\alpha(u)+\alpha^{2}(v)$, where $u \in T_{x_{1}}$ and $v \in T_{x_{2}}$ and they are $(n-1)^{2}$ in number. We write all such elements in index row and column and complete the multiplication table of $T_{x_{1}} \oplus T_{x_{2}}$. We see that no new element appear in the table. Then the number of elements in the index row or column is $2 n-1+(n-1)^{2}=n^{2}$. We now consider $n=3$. Take the multiplication table of $T_{x_{1}} \oplus T_{x_{2}}$, and write all elements of $T_{x_{3}}$ except 0 in the index row and column. The number of elements in the index row and column are $n^{2}+n-1$. Multiply the elements of $T_{x_{1}} \oplus T_{x_{2}}$ and $T_{x_{3}}$. Then in the table, some new elements of the form $t \circ w=\alpha(t)+\alpha^{2}(w)$ appear, where $t \in T_{x_{1}} \oplus T_{x_{2}}, w \in T_{x_{3}}$ which are $n^{2}(n-1)$ in number. Now we write all these elements in the index row and column of the table of $T_{x_{1}} \oplus T_{x_{2}} \oplus T_{x_{3}}$. We see that no new element appears in the table. The number of elements in the index row or column is $n^{2}+n^{2}(n-1)=n^{3}$. Continuing in this way we finally get $\left|T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{r}}\right|=n^{r}$.

Theorem 10. Let $p$ be prime and $F_{P}$ a finite field. Let $E$ denote the $r$-th extension of $F_{P}$. Then there exists a unique epimorphism between LAsemigroups formed by $E$ and $F_{p}$.

Proof. Let $\alpha$ be a unary operation. Suppose that $\beta$ is a root of an irreducible polynomial with respect to $F_{p}$. It is not difficult to prove that the mapping
$\varphi: E \rightarrow F_{P}$ defined by $\varphi\left(a_{0}+a_{1} \beta+\ldots+a_{r-1} \beta^{r-1}\right)=a_{0}+a_{1}+\ldots+a_{r}$ is a unique epimorphism.

Theorem 11. $T_{x}$ is simple.
Proof. Suppose that $T_{x}$ has a proper left (right) ideal of $S$. Then by definition $S T_{x} \subseteq S\left(T_{x} S \subseteq S\right)$ and $S$ is proper LA-subsemigroup of $T_{x}$. We know that the order of $T_{x}$ is either prime or power of a prime. So, if it has a proper LA-subsemigroup $S$, then the order of $S$ will be prime. Since $S$ is embedded into $T_{x}$, so there exists a monomorphism between $T_{x}$ and $S$. But by Theorem 10, there exists a unique epimorphism between $T_{x}$ and $S$. This implies that there exists an isomorphism between $T_{x}$ and $S$. This is a contradiction. Hence the proof.

Theorem 12. If $K$ is a kernel of a homomorphism $h$ between LA-groups $W$ and $W^{\prime}$, then
(a) $K \leq W$,
(b) $W / K$ is an LA-group,
(c) $W / K \cong \operatorname{Im}(h)$.

Proof. (a) and (b) are obvious. For (c) define a mapping $\varphi: W / K \rightarrow$ $\operatorname{Im}(h)$ by $\varphi(u \circ K)=h(u)$ for $u \in W$. Then $\varphi$ is an isomorphism.

Theorem 13. If $T_{1}=T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{n}}, \quad T_{2}=T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{m}}$, where $n \neq m$, then
(1) $T_{1} \leq T_{1} \oplus T_{2}$ and $T_{1} \cap T_{2} \leq T_{2}$,
(2) $T_{1} \oplus T_{2} / T_{1}$ and $T_{2} / T_{1} \cap T_{2}$ are $L A$-semigroups,
(3) $T_{1} \oplus T_{2} / T_{1} \cong T_{2} / T_{1} \cap T_{2}$.

Proof. (1) and (2) are obvious. For (3) define a mapping $\varphi: T_{2} / T_{1} \cap T_{2} \longrightarrow$ $T_{1} \oplus T_{2} / T_{1}$ by $\varphi\left(v \circ\left(T_{1} \cap T_{2}\right)\right)=v \circ T_{1}$ for all $v \in T_{1} \cap T_{2}$. Then $\phi$ is an isomorphism.

Theorem 14. If $W_{X}$ is an LA-group, and $T=T_{x_{1}} \oplus T_{x_{2}} \oplus \ldots \oplus T_{x_{n}}$, then $\left(W_{X} / T_{x_{i}}\right) /\left(T / T_{x_{i}}\right)$ is isomorphic to $W_{X} / T$, where $1 \leqslant i \leqslant n$.

Proof. Define a mapping $\varphi: W_{X} / T_{x_{i}} \longrightarrow W_{X} / T$, by $\varphi\left(v \circ T_{x_{i}}\right)=v \circ T$, where $v \in W_{X}$. Then $\varphi$ is an epimorphism. By Theorem 12,

$$
\left(W_{X} / T_{x_{i}}\right) /(\operatorname{Ker} \varphi) \cong W_{X} / T
$$

and $\operatorname{Ker} \varphi=T / T_{x_{i}}$. Hence the proof.

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# Quasi union hyper K-algebras 

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#### Abstract

We give a method of construction of a hyper K-algebra on a set of order $\alpha$, where $\alpha$ is a fixed cardinal number. Then we introduce the notion of quasi union hyper K-algebra and prove that any quasi union hyper K-algebra is implicative and whenever $0 \circ 0=\{0\}$, it is strong implicative hyper Kalgebra. Also a quasi union hyper K -algebra is positive implicative if and only if it is a hyper BCK-algebra. Finally we prove that any hyper Kalgebra $H \stackrel{\mathrm{C}}{=} \oplus_{i \in \Lambda} A_{i}$ (closed set), where $\left|A_{i}\right|=2$ under some conditions is a quasi union hyper K-algebra or a quasi union hyper BCK-algebra.


## 1. Introduction

The study of BCK-algebra was initiated by Imai and Iséki [6] in 1966 as a generalization of the concept of set-theoretic difference and propositional calculi. The hyper structure theory (called also multi algebras) was introduced in 1934 by Marty [8] at the 8th congress of Scandinavian Mathematicians. Hyper structures have many applications to several sectors of both pure and applied sciences. Borzooei, et.al. [4, 7] applied the hyper structure to BCK-algebras and introduced the concept of hyper BCKalgebra and hyper K-algebra in which, each of them is a generalization of BCK-algebra. Borzooei and Harizavi [3] introduced a decomposition for a hyper BCK-algebra. Nasr-Azadani and Zahedi [9] study S-absorbing (P)decomposable hyper K-algebras as a generalization of decomposition for hyper BCK-algebras. Now, we follow [9] and obtain some results as mentioned in the abstract.

[^6]
## 2. Preliminaries

Let $H$ be a non-empty set, the set of all non-empty subset of H is denoted by $\mathcal{P}^{*}(H)$. A hyperoperation on $H$ is a map $\circ: H \times H \rightarrow \mathcal{P}^{*}(H)$, where $(a, b) \rightarrow a \circ b$ for all $a, b \in H$. A set $H$, endowed with a hyperoperation, "०", is called a hyperstructure. If $A, B \subseteq H$, then $A \circ B=\bigcup_{a \in A, b \in B} a \circ b$.

Definition 1. $[4,7]$ Let H be a non-empty set containing a constant "0" and "०" be a hyperoperation on $H$. Then H is called a hyper $K$-algebra (hyper BCK-algebra) if it satisfies K1 - K5 (respectively: HK1 - HK4).

| K1: $(x \circ z) \circ(y \circ z)<x \circ y$, | HK1: $(x \circ z) \circ(y \circ z) \ll x \circ y$, |
| :--- | :--- |
| K2: $(x \circ y) \circ z=(x \circ z) \circ y$, | HK2: $(x \circ y) \circ z=(x \circ z) \circ y$, |
| K3: $x<x$, | HK3: $x \circ H \ll x$, |
| K4: $x<y, y<x$, then $x=y$, | HK4: $x \ll y, y \ll x$, then $x=y$, |
| K5: $0<x$ |  |

for all $x, y, z \in H$, where $x<y(x \ll y)$ means $0 \in x \circ y$. Moreover for any $A, B \subseteq H, A<B$ if $\exists a \in A, \exists b \in B$ such that $a<b$ and $A \ll B$ if $\forall a \in A, \exists b \in B$ such that $a \ll b$.

For briefly the readers could see some definitions and results about hyper K-algebra and hyper BCK-algebra in [4, 7]. In the sequel $H$ always denotes a hyper K-algebra. If $I \subset H$, then $I^{\prime}=H \backslash I$ and $I^{*}=I^{\prime} \cup\{0\}$.

Definition 2. [5] An element $b \in H$ is called a left (right) scalar if $|b \circ x|=1$ $(|x \circ b|=1)$ for all $x \in H$. An element is called scalar if it is a left and a right scalar.

Theorem 1. [10] Let $\left(H_{i}, \circ_{i}, 0\right), i \in \Omega$ be a family of hyper $K$-algebras such that $H_{i} \cap H_{j}=\{0\}, i \neq j \in \Omega, 0$ be a left scalar in each $H_{i}, i \in \Omega$, $H=\bigcup_{i \in \Omega} H$ and "०" on $H$ is defined as follows:

$$
x \circ y:= \begin{cases}x \circ_{i} y & \text { if } x, y \in H_{i}, \\ \{x\} & \text { if } x \in H_{i}, y \notin H_{i} .\end{cases}
$$

Then $(H, \circ, 0)$ is hyper $K$-algebra denoted by $H=\oplus_{i \in \Omega} H_{i}$.
Definition 3. [1, 2] A hyper K-algebra $H$ is called
(i) weak implicative if $x<x \circ(y \circ x)$,
(ii) implicative if $x \in x \circ(y \circ x)$,
(iii) strong implicative if $x \circ 0 \subseteq x \circ(y \circ x)$,
(iv) positive implicative if $(x \circ y) \circ z=(x \circ z) \circ(y \circ z)$
holds for all $x, y, z \in H$.

Definition 4. [9, 4, 11] A non-empty subset $I$ of $H$ is said to be closed if $x<y$ and $y \in I$ imply $x \in I$, and it is said to be a hyper $K$-ideal of $H$ if $x \circ y<I$ and $y \in I$ imply $x \in I$.

Theorem 2. [9] Any hyper $K$-ideal of $H$ is closed.
Definition 5. [9] Let $I$ and $S$ be non-empty subsets of $H$. Then we say that $I$ is $S$-absorbing if $x \in I$ and $y \in S$ imply $x \circ y \subseteq I$. In the case $S=I^{\prime}$ or $S=I^{*}$ we say that $I$ is $C$-absorbing or $C^{*}$-absorbing, respectively.

Theorem 3. [9] Let H be a hyper BCK-algebra and I be a hyper BCK-ideal or closed set. Then $I$ is $H$-absorbing.

Definition 6. [9] A hyper K-algebra $H$ is called ( $P$ )-decomposable if there exists a non-trivial family $\left\{A_{i}\right\}_{i \in \Lambda}$ of subsets of $H$ with $P$-property such that $H \neq\left\{A_{i}\right\}$ for all $i \in \Lambda, H=\bigcup_{i \in \Lambda} A_{i}$ and $A_{i} \cap A_{j}=\{0\}, i \neq j$.

In this case, we write $H=\oplus_{i \in \Lambda} A_{i}(\mathrm{P})$ and say that $\left\{A_{i}\right\}_{i \in \Lambda}$ is a (P)decomposition for $H$. If each $A_{i}, i \in \Lambda$, is S -absorbing we write $H \stackrel{\mathrm{~S}}{=}$ $\oplus_{i \in \Lambda} A_{i}(\mathrm{P})$. Moreover, we say that this decomposition is closed union, in short (P)-CUD, if $\cup_{i \in \Delta} A_{i}$ has P-property for any non-empty subset $\Delta$ of $\Lambda$. If there exists a (P)-CUD for $H$, then we say that $H$ is a (P)-CUD.

Theorem 4. [9] Let $H \stackrel{H}{=} A \oplus B$. Then 0 is a left scalar element.
Theorem 5. [9] Let $H \stackrel{\mathrm{C}^{*}}{=} \oplus_{i \in \Lambda} A_{i}$ (hyper K-ideal). Then $H$ is (hyper $K$ -ideal)-CUD and $H \stackrel{\mathrm{C}^{*}}{=} I \oplus I^{*}\left(\right.$ hyper $K$-ideal), where $I=\cup_{i \in \Delta} A_{i}$ for any non-empty subset $\Delta$ of $\Lambda$.

Theorem 6. [10] Let $(H, \circ, 0)$ be a hyper BCK-algebra. Then $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-algebra) if and only if $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-ideal).

## 3. Quasi union hyper K-algebra

In this section we give a method to construct a hyper K-algebra of order $\alpha$ where $\alpha$ is a given cardinal number. Also we introduce the concept of quasi union hyper K-algebra and investigate some properties of it.

Remark 1. Let $H$ be a set containing " 0 ", $\mathcal{P}_{0}(H)=\{A \subseteq H: 0 \in A\}$ and $\mathcal{S}=\left\{f \mid f: H \rightarrow \mathcal{P}_{0}(H)\right.$ is a function $\}$. For convenience we use $F^{x}$ instead of $f(x)$ for any $f \in \mathcal{S}$. Clearly $\mathcal{S} \neq \emptyset$, because the functions $f, g: H \rightarrow \mathcal{P}_{0}(H)$, where $f(x)=\{0\}$ and $g(x)=\{0, x\}$ for all $x \in H$, are members of $\mathcal{S}$.

Theorem 7. Let $H=X \cup\{0\}$, where $X$ is a non-empty set. Then for any $f \in \mathcal{S}$ we can define the hyperoperation $\circ_{f}: H \times H \longrightarrow \mathcal{P}^{*}(H)$ by putting:

$$
x \circ_{f} y:= \begin{cases}F^{x} & \text { if } x=y, \\ \{x\} & \text { otherwise } .\end{cases}
$$

Moreover, the following statements are equivalent
(i) $\left(H, \circ_{f}, 0\right)$ is a hyper $K$-algebra,
(ii) $F^{x} \circ_{f} y=F^{x}$ for all $y \neq x, y \in H$,
(iii) $x \neq y$ and $y \in F^{x}$ imply $y \in F^{y}$ and $F^{y} \subseteq F^{x}$.

Proof. By Remark 1, $u=v$ implies $f(u)=F^{u}=f(v)=F^{v}$. This yields that " $\circ_{f}$ " is well-defined and hence it is a hyperoperation on $H$.
$(i) \Rightarrow(i i)$. Let $\left(H, \circ_{f}, 0\right)$ be a hyper K-algebra and $y \neq x, y \in H$. Then by definition of " $\circ_{f}$ " and K2 we have:

$$
F^{x} \circ_{f} y=\left(x \circ_{f} x\right) \circ_{f} y=\left(x \circ_{f} y\right) \circ_{f} x=\left(x \circ_{f} x\right)=F^{x} .
$$

$(i i) \Rightarrow(i)$. To do this, we show that $H$ satisfies K1 - K5. Since $0 \in$ $F^{x}=x \circ_{f} x$, hence $x<x$ for all $x \in H$ and K3 holds. Moreover by definition of $\circ_{f}$ we have $0 \circ_{f} x=\{0\}$ for all $x \neq 0$, that is $0<x$. Thus K5 holds.

To check K1, K2 and K4, we consider the following five cases:

$$
\begin{array}{ll}
\text { (I) } x \neq y, x \neq z \text { and } y \neq z, & \text { (II) } x=y \neq z, \\
\text { (IV) } x \neq y=z, & \text { (III) } x=z \neq y, \\
\text { (V) } x=y=z . &
\end{array}
$$

K1: $\left(x \circ_{f} z\right) \circ_{f}\left(y \circ_{f} z\right)<x \circ_{f} y$.
For convenience, we put $\left(x \circ_{f} z\right) \circ_{f}\left(y \circ_{f} z\right)=A$ and $x \circ_{f} y=B$. If (I) holds, then $A=\{x\}=B$ and by K3, $A<B$. If (II) holds, then $A=F^{x}=B$, therefore $A<B$. If (III) holds, then by (ii), $A=F^{x}{ }_{\circ} y=F^{x}$ and $B=\{x\}$. Since $0 \in F^{x}$ and K5 holds, then $A<B$. If (IV) holds, then $A=x \circ_{f} F^{y}$ and $B=\{x\}$. Since $0 \in F^{y}$ and K3 holds, thus $x \in x \circ_{f} 0 \subseteq x \circ_{f} F^{y}$ and it yields that $A<B$. If (V) holds, then $A=F^{x} \circ_{f} F^{x}$ and $B=F^{x}$. Since $0 \in F^{x}$ and K5 holds, then $A<B$. Therefore K1 holds in all cases.

K2: $\left(x \circ_{f} y\right) \circ_{f} z=\left(x \circ_{f} z\right) \circ_{f} y$.
We put $A=\left(x \circ_{f} y\right) \circ_{f} z$ and $B=\left(x \circ_{f} z\right) \circ_{f} y$ and show that $A=B$ for all cases (I) $-(\mathrm{V})$. If (I) holds, then $A=\{x\}=B$. If (II) holds, then by (ii) we have $A=F^{x} \circ_{f} z=F^{x}$ and $B=F^{x}$, so $A=B$. If (III) holds, similar to the proof of case (II) we have $A=B$. If (IV) holds, then $A=\{x\}=B$. If (V) holds, then $A=B$. Finally we show that K4 holds, i.e., $x<y, y<x \Rightarrow x=y$. Suppose $x<y, y<x$ and $x \neq y$. Then we
have $0 \in x \circ_{f} y=\{x\}$ and $0 \in y \circ_{f} x=\{y\}$. Hence $x=y=0$ which is a contradiction to $x \neq y$. Thus ( $\mathrm{H}, \circ_{f}, 0$ ) is hyper K-algebra.
(ii) $\Rightarrow(i i i)$. Let $y \neq x$ and $y \in F^{x}$. Then, according to the definition, $u \circ_{f} y=\{u\}$ where $u \neq y$. Therefore

$$
\begin{equation*}
F^{x} \circ_{f} y=\cup_{u \neq y, u \in F^{x}}\left(u \circ_{f} y\right) \cup y \circ y=\left(F^{x}-\{y\}\right) \cup F^{y} . \tag{1}
\end{equation*}
$$

By (ii), $F^{x} \circ_{f} y=F^{x}$. So equality (1) yields that $y \in F^{y}$ and $F^{y} \subseteq F^{x}$, that is, (iii) holds.
(iii) $\Rightarrow$ (ii). Suppose $x \neq y$. We consider two cases $(a): y \notin F^{x}$ and (b): $y \in F^{x}$. If (a) holds, then $u \neq y$ for all $u \in F^{x}$. Thus by definition of $\circ_{f}$ we have $F^{x} \circ_{f} y=F^{x}$, hence (ii) holds. If ( $b$ ) holds, then by equality (1) and hypothesis $\left(F^{y} \subseteq F^{x}\right)$ we get that $F^{x} \circ_{f} y=F^{x}$.

Definition 7. The hyperoperation and hyper K-algebra which have been introduced in Theorem 7 are called a quasi union hyper operation and a quasi union hyper $K$-algebra, respectively.
Corollary 1. For any set $X$ such that $0 \notin X$ and $f(x) \in\{\{0\},\{0, x\}\}$ for all $f \in \mathcal{S}$ and $x \in H$ there is a quasi union hyper $K$-algebra on $H=X \cup\{0\}$ with the hyperoperation defined as follows:

$$
x \circ y:= \begin{cases}F^{x}=\{0\} \text { or } F^{x}=\{0, x\} & \text { if } x=y \\ \{x\} & \text { otherwise } .\end{cases}
$$

Proof. Since $F^{x} \circ y=F^{x}$, for all $x \neq y \in H$, thus by Theorem 7 (ii) and Definition $7,(H, \circ, 0)$ is a quasi union hyper K-algebra.

Example 1. Let $X=\{1,2\}$. Then according to Corollary 1, each of the following tables are quasi union hyper K-algebra on $H=\{0,1,2\}$.

| $\circ_{1}$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0\}$ |


| $\circ_{2}$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0\}$ |


| $\circ_{3}$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,2\}$ |


| $\circ_{4}$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0,1\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,1,2\}$ |

Corollary 2. Let $H$ be a quasi union hyper $K$-algebra and $x \neq y$. If $y \in F^{x}$ and $x \in F^{y}$, then $F^{y}=F^{x}$.

The proof follows from Theorem 7 (iii).

## 4. Some results on quasi union hyper K-algebras

Theorem 8. Let $H$ be a quasi union hyper K-algebra. Then the following statements are equivalent:
(i) $H$ is positive implicative hyper $K$-algebra,
(ii) $F^{x}=\{0\}$ or $F^{x}=\{0, x\}$ for all $x \in H$,
(iii) $H$ is a hyper BCK-algebra.

Proof. $(i) \Rightarrow(i i)$. Let $H$ be positive implicative, i.e., $(x \circ y) \circ z=(x \circ z) \circ(y \circ$ z) for all $x, y, z \in H$ and $u \in F^{x}$. If $u \neq x$, since $(u \circ x) \circ x=(u \circ x) \circ(x \circ x)$ we get that $\{u\}=\{u\} \circ(x \circ x)$. From $u \in F^{x}=x \circ x$, we conclude that $0 \in\{u\} \circ(x \circ x)=\{u\}$. So $u=0$ and $F^{x}=\{0\}$ or $F^{x}=\{0, x\}$ for all $x \in H$.
(ii) $\Rightarrow(i)$. Suppose $F^{x}=\{0\}$ or $F^{x}=\{0, x\}$ for all $x \in H$. We show that $H$ is a positive implicative hyper K-algebra, i.e., $H$ satisfies the following identity:

$$
\begin{equation*}
(x \circ y) \circ z=(x \circ z) \circ(y \circ z) . \tag{2}
\end{equation*}
$$

We prove it by considering the following cases: (I) $x \circ x=\{0\}$, (II) $x \circ x=\{0, x\}$.
CASE 1. $x \neq y, x \neq z, y \neq z$.

$$
(x \circ y) \circ z=\{x\} \circ z=\{x\} \text { and }(x \circ z) \circ(y \circ z)=\{x\} \circ\{y\}=\{x\} .
$$

CASE 2. $x=y \neq z$. If (I) holds, then

$$
(x \circ y) \circ z=\{0\} \circ z=\{0\} \text { and }(x \circ z) \circ(y \circ z)=\{x\} \circ\{x\}=\{0\} .
$$

If (II) holds, then

$$
(x \circ y) \circ z=\{0, x\} \circ z=\{0, x\} \text { and }(x \circ z) \circ(y \circ z)=\{x\} \circ\{x\}=\{0, x\} .
$$

Case 3. $x=z \neq y$. By K2 and the proof of Case 2, (2) holds.
CASE 4. $x \neq y=z$. By considering $F^{0}=0 \circ 0=\{0\}$, if ( I ) holds then

$$
(x \circ y) \circ z=\{x\} \circ z=\{x\} \text { and }(x \circ z) \circ(y \circ z)=\{x\} \circ\{0\}=\{x\} .
$$

If (II) holds, then

$$
(x \circ y) \circ z=\{x\} \circ z=\{x\} \text { and }(x \circ z) \circ(y \circ z)=\{x\} \circ\{0, y\}=\{x\} .
$$

CASE 5. $x=y=z$. By considering $F^{0}=0 \circ 0=\{0\}$, if (I) holds then

$$
(x \circ y) \circ z=\{0\} \circ x=\{0\} \text { and }(x \circ z) \circ(y \circ z)=\{0\} \circ\{0\}=\{0\} .
$$

If (II) holds, then $(x \circ y) \circ z=\{0, x\} \circ x=\{0, x\}$ and $(x \circ z) \circ(y \circ z)=$ $\{0, x\} \circ\{0, x\}=\{0, x\}$. These cases imply that the identity (2) is satisfied, thus $H$ is a positive implicative hyper K -algebra.
$(i i) \Rightarrow(i i i)$. Let $F^{x}=\{0\}$ or $F^{x}=\{0, x\}$ for all $x \in H$. We show that $H$ is a hyper BCK-algebra. To do this, since each hyper K-algebra satisfies HK2 and HK4, it is sufficient to prove $H$ satisfies HK1 and HK3. Now we show that HK1 holds, i.e., $(x \circ z) \circ(y \circ z) \ll x \circ y$ for all $x, y \in H$. We prove it by considering the following cases:

$$
\text { (I) } x \circ x=\{0\}, \quad \text { (II) } x \circ x=\{0, x\} \text {. }
$$

Case 1. $x \neq y, x \neq z, y \neq z$.

$$
(x \circ z) \circ(y \circ z)=\{x\} \ll x \circ y=\{x\} .
$$

CASE 2. $x=y \neq z$.

$$
(x \circ z) \circ(y \circ z)=\{x\} \circ\{x\}=x \circ x \ll x \circ y=x \circ x .
$$

Case 3. $x=z \neq y$. By considering $F^{0}=0 \circ 0=\{0\}$, if (I) holds then

$$
(x \circ z) \circ(y \circ z)=\{0\} \circ\{y\}=\{0\} \ll x \circ y=\{x\} .
$$

If (II) holds, then $(x \circ z) \circ(y \circ z)=\{0, x\} \circ\{y\}=\{0, x\} \ll x \circ y=\{x\}$.
CASE 4. $x \neq y=z$. If (I) holds, then $(x \circ z) \circ(y \circ z)=\{x\} \circ\{0\}=\{x\} \ll\{x\}$. If (II) holds, then $(x \circ z) \circ(y \circ z)=\{x\} \circ\{0, y\}=\{x\} \ll x \circ y=\{x\}$.
CASE 5. $x=y=z$. If (I) holds, then $(x \circ z) \circ(y \circ z)=\{0\} \ll x \circ y=\{0\}$. If (II) holds, then $(x \circ z) \circ(y \circ z)=\{0, x\} \ll x \circ y=\{0, x\}$.

Therefore HK1 holds. Finally since $0 \ll x, x \ll x$, hence $\{0, x\} \ll x$. Therefore by considering "०" of $H$ we have $x \circ y \ll x$ for all $x, y \in H$, i.e., HK3 holds. Thus $H$ is a hyper BCK-algebra.
$(i i i) \Rightarrow(i i)$. Let $H$ be a quasi union hyper BCK-algebra. Then $F^{0}=$ $0 \circ 0=\{0\}$. So, let $u \in F^{x}$ and $u \neq x$. Then, since $x \circ x \ll x$, we have $u \ll x$ or $0 \in u \circ x=\{u\}$. This implies that $u=0$, hence $F^{x}=\{0\}$ or $F^{x}=\{0, x\}$ for all $x \in H$.

Theorem 9. Any quasi union hyper $K$-algebra $H$ is implicative.
Proof. Let $H$ be a quasi union hyper K-algebra. By considering Definition 3, it is enough to show that $x \in x \circ(y \circ x)$ for all $x, y \in H$. Let $x, y \in H$. Then if $x \neq y$, we have $x \circ(y \circ x)=\{x\}$ and if $x=y$, then $x \in x \circ(x \circ x)$. Because $0 \in x \circ x$. Hence we have $x \in x \circ(y \circ x)$, for any $x, y \in H$.

Theorem 10. Let $H$ be a quasi union hyper K-algebra. Then $H$ is strong implicative if and only if $F^{0}=\{0\}$.

Proof. Let $H$ be a strong implicative quasi union hyper K-algebra. Then $x \circ 0 \subseteq x \circ(y \circ x)$ for all $x, y \in H$. If $x=0$ and $y \neq 0$ we have $0 \circ 0 \subseteq$ $0 \circ(y \circ 0)=\{0\}$. Hence $0 \circ 0=F^{0}=\{0\}$. Conversely, suppose $F^{0}=\{0\}$. We prove that $x \circ 0 \subseteq x \circ(y \circ x)$ for all $x, y \in H$. By considering $F^{0}=0 \circ 0=\{0\}$, if $x \neq y$, then we have $x \circ 0=\{x\}=x \circ(y \circ x)$. If $x=y$, then we have $x \circ 0=\{x\} \subseteq x \circ(x \circ x)$, because $0 \in x \circ x$ and $x \circ 0=\{x\}$. Therefore $H$ is a strong implicative hyper K-algebra.

Theorem 11. If $(H, \circ, 0)$ is a quasi union hyper $K$-algebra, then for any $x \in H \backslash\{0\}, A_{x}=\{0, x\}$ is a hyper $K$-ideal of $H$.

Proof. Suppose $v \circ y<A_{x}$ and $y \in A_{x}$. We show that $v \in A_{x}$. If $v \in\{0, x\}$, then we are done. Otherwise, we have $v \circ y=\{v\}<\{0, x\}$. This implies that $v<0$ or $v<x$. Since $v \neq 0, x$, from these we conclude that $0 \in v \circ 0=\{v\}$ or $0 \in v \circ x=\{v\}$. Hence $v=0$, which is a contradiction. Therefore $v \in\{0, x\}$ and hence $A_{x}$ is a hyper K-ideal of $H$.

Theorem 12. Let $H$ be a quasi union hyper $K$-algebra. Then $H \stackrel{\mathrm{C}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper K-ideal).

Proof. By considering Definition 6 and Theorem 11, it is enough to show that for all $x \in H \backslash\{0\}, A_{x}=\{0, x\}$ is C-absorbing. Suppose $t \notin\{0, x\}$, since $u \circ t=\{u\} \subseteq A_{x}$ for all $u \in\{0, x\}$, we conclude that $A_{x}$ is Cabsorbing.

Corollary 3. Let $H$ be a quasi union hyper $K$-algebra and $0 \circ 0=\{0\}$. Then $H \stackrel{\mathrm{C}^{*}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}($ hyper $K$-ideal).

Proof. The proof follows from Definition 5 and Theorem 12.
By the following example we show that there is a quasi union hyper K-algebra such that $A_{x}=\{0, x\}$ is not $C^{*}$-absorbing.

Example 2. Consider $H=\{0,1,2\}$ with the following structure:

| $\circ$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0,1\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,1,2\}$ |

Then $(H, 0,0)$ is a quasi hyper K -algebra and $A_{2}=\{0,2\}$ is not $C^{*}$ absorbing, because $0 \circ 0=\{0,1\} \nsubseteq A_{2}$.

Corollary 4. Let $H$ be a quasi union hyper $K$-algebra. Then $H \stackrel{\mathrm{C}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (closed set).

Proof. Since any hyper K-ideal is closed set, the proof follows from Theorem 12.

Lemma 1. Any hyper K-ideal I of hyper BCK-algebra $H$ is a hyper BCKideal too.

Proof. Let $x \circ y \ll I$ and $y \in I$. Then $x \circ y<I$. Since $I$ is a hyper K-ideal and $y \in H$, we conclude that $x \in I$. Hence $I$ is a hyper BCK-ideal of $H$.

Corollary 5. Let $H$ be a quasi union hyper BCK-algebra. Then $H \stackrel{H}{=}$ $\oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper BCK-ideal).

Proof. Since by Theorem 3 any hyper BCK-ideal is H-absorbing, then by using Lemma 1 and Theorem 12 we get that $H \stackrel{H}{=} \oplus_{x \in H}\{0, x\}$ ( hyper BCKideal).

Corollary 6. Let $H$ be a quasi union hyper BCK-algebra. Then $H \stackrel{H}{=}$ $\oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper BCK-algebra), i.e., it is a union of family of hyper BCKalgebras.

Proof. The proof follows from Corollary 5 and Theorem 6.
Theorem 13. Any quasi union hyper $K$-algebra $H$ is (hyper $K$-ideal)-CUD.
Proof. By Theorem 12, $H \stackrel{\text { C }}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper K-ideal). By Theorem 1, we must show that for any non-empty subset $B$ of $H \backslash\{0\}, \cup_{x \in B} A_{x}$ is a hyper K-ideal of $H$. Suppose $u \circ y<\cup_{x \in B} A_{x}$ and $y \in \cup_{x \in B} A_{x}$. If $u \neq y$ then $u \circ y=\{u\}<\cup_{x \in B} A_{x}$. This yields that for some $x \in B, u<A_{x}$. Since $A_{x}$ is a hyper K-ideal and by Theorem 2 it is a closed set, we conclude that $u \in A_{x}$. Therefore $u \in \cup_{x \in B} A_{x}$. If $u=y$, then $u \in \cup_{x \in B} A_{x}$. Thus $\cup_{x \in B} A_{x}$ is a hyper K -ideal of $H$, i.e., $H$ is a (hyper K-ideal)-CUD.

Theorem 14. Let $H$ be a quasi union hyper $K$-algebra and I be a subset of $H$ containing 0. Then $I$ is a hyper $K$-ideal of $H$.

Proof. By Theorem 12 we have $H \stackrel{\mathrm{C}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper K-ideal). Since $I=\cup_{x \in I}\{0, x\}$, by Theorem 13, $I$ is a hyper K-ideal of $H$.

Now, we proceed to find some relations between a quasi union hyper K-algebra and a family of hyper K-algebras of type $H \stackrel{\mathrm{C}}{=} \oplus_{i \in \Lambda} A_{i}$ (hyper Kideal) where, $\left|A_{i}\right|=2$. In particular, we show that whenever $|H| \geqslant 4$, any type of these hyper K-algebras is a quasi union hyper K-algebra.
Remark 2. Let $H \stackrel{\mathrm{C}}{=} \oplus_{i \in \Lambda} A_{i}$ (hyper K-ideal) where, $\left|A_{i}\right|=2$. Since $\left|A_{i}\right|=$ 2, we have $A_{i}=\{0, x\}$ for a nonzero element $x \in H$. Hence for convenience we write $A_{x}$ instead of $A_{i}$ and hence $H \stackrel{\mathrm{C}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper K-ideal).

Theorem 15. Let $H \stackrel{\mathrm{C}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper $K$-ideal) and $|H| \geqslant 4$. Then $H$ is a quasi union hyper $K$-algebra.
Proof. Since by K3, we have $0 \in x \circ x=F^{x}$, according to Theorem 7, it is sufficient to show that $x \circ y=\{x\}$ for all $x \neq y$. Suppose $u \in x \circ y$ and $x \neq y$. Then by considering the following three cases we prove $u=x$.

$$
\text { (I) } y=0, \quad \text { (II) } x \neq 0 \text { and } y \neq 0, \quad \text { (III) } x=0 .
$$

If (I) holds, then since $x \circ 0<A_{u}$ and $A_{u}$ is a hyper K-ideal, we conclude that $x \in A_{u}$. Since $x \neq y=0$, then $x=u$. If (II) holds, since $y \notin A_{x}$ and $A_{x}$ is C-absorbing, we get that $x \circ y \subseteq A_{x}$. Thus $u \in A_{x}$. We show that $u \neq 0$. If $u=0$, then $x<y$ and $x \in A_{y}$, because any hyper K-ideal is closed set. This yields that $x=y$, which is a contradiction. Therefore $u=x$. If (III) holds, then since $|H| \geqslant 4$ we have at least two nonzero elements $t, z \in H$ different from $y$. Therefore $0 \circ y \subseteq A_{t} \cap A_{z}=\{0\}$, because $A_{x}$ and $A_{t}$ are C-absorbing. This yield that $0 \circ y=\{0\}$, or $u=x=0$. Therefore $x \circ y=\{x\}$, where $x \neq y \in H$.

Theorem 15 is not true in general.
Example 3. Let $H=\{0,1,2\}$ with the following structure:

| $\circ$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0,2\}$ | $\{0,1\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,2\}$ |

Then $H=(H, \circ, 0)$ is a hyper K-algebra such that $H \stackrel{\mathrm{C}}{=}\{0,1\} \oplus\{0,2\}$ (hyper K-ideal) and $0 \circ y \neq\{0\}$ where $y \neq 0$. Also this example shows that even if each $A_{x}$ in Theorem 15 is $C^{*}$-absorbing, then $H$ may not be a quasi union hyper K-algebra, whenever $|H|=3$.

Lemma 2. Let $H \stackrel{H}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (closed set) and $|H| \geqslant 3$. Then 0 is a left scalar.

Proof. Since $|H| \geqslant 3$ the proof follows from Theorems 5 and 4.
Theorem 16. Let $H \stackrel{H}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (closed set) and $|H| \geqslant 3$. Then $x \circ y$ $=\{x\}$ for $x \neq y$.

Proof. By Lemma 2 we conclude that $0 \circ y=\{0\}$ for all $y \in H$. Now let $0 \neq x \neq y$. On the contrary, suppose $x \circ y \neq\{x\}$. Since $A_{x}$ is H -absorbing we have $x \circ y \subseteq A_{x}=\{0, x\}$. If $x \circ y=\{0, x\}$ or $\{0\}$, then $x<y$. In this case if $y=0$ we conclude that $x=0$, which is a contradiction. Otherwise, $y \neq 0$, we get that $x \in A_{y}$, because $A_{y}$ is a closed set and $y \in A_{y}$. This yields that $x=y$ which is also a contradiction. Hence $x \circ y=\{x\}$. So, $x \neq y$.

Theorem 16 is not true in general.
Example 4. Let $H=\{0,1\}$ with the following structure:

| $\circ$ | 0 | 1 |
| :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0,1\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ |

Then $H=(H, \circ, 0)$ is a hyper K -algebra such that $0 \circ 1 \neq\{0\}$.
Theorem 17. Let $H \stackrel{H}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (closed set) and $|H| \geqslant 3$. Then 0 is a scalar and $x \circ y=\{x\}$ for $x \neq y$.

Proof. By Theorem 16, $a \circ 0=\{a\}$ and $0 \circ a=\{0\}$ while $a \neq 0$. Also by Lemma 2 we have $0 \circ 0=\{0\}$. Hence 0 is scalar. The remaining of the proof follows from Theorem 16.

Corollary 7. Let $H \stackrel{\mathrm{H}}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (hyper K-ideal) and $|H| \geqslant 3$. Then 0 is a scalar and $x \circ y=\{x\}$ for $x \neq y$.

The proof follows from Theorems 2 and 17.
Theorem 18. Let $H \stackrel{H}{=} \oplus_{x \in H \backslash\{0\}} A_{x}$ (closed set) and $|H| \geqslant 3$. Then $H$ is a positive (strong) implicative quasi union hyper BCK-algebra.

Proof. By hypothesis and Theorem 17, we have $0 \circ 0=\{0\}$ and $x \circ y=\{x\}$, where $x \neq y$. Since $A_{x}$ is H -absorbing we have $x \circ x \subseteq A_{x}$, for all $x \in H$. Hence $x \circ x=\{0\}$ or $x \circ x=\{0, x\}$. Therefore these imply that

$$
x \circ y= \begin{cases}\{0\} \text { or }\{0, x\} & \text { if } x=y, \\ \{x\} & \text { otherwise } .\end{cases}
$$

So the proof follows from Corollary 1 and Theorems 8 and 10 .

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# A note on a union of hyper K-algebras 

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#### Abstract

In this paper, at first by some examples we show that the Theorem 3.7 of [1] is not true in general. Then we give a correct version of it. Moreover we give the notion of a union of a family of hyper K-algebras and investigate some of its properties. Finally by considering the concept of (closed set)-decomposable, we show that a hyper BCK-algebra is (closed set)-decomposable if and only if it is the union of a family of hyper BCKalgebras.


## 1. Introduction and Preliminaries

The study of BCK-algebra was initiated by Imai and Iséki [2] in 1966 as a generalization of the concept of set-theoretic difference and propositional calculi. Borzooei, et.al. [1, 3] applied the hyper structure to BCK-algebras and introduced the concept of hyper BCK-algebra and K-algebra in which each of them is a generalization of BCK-algebra. They have defined the notion of a union of two hyper K-algebra as an extension of a union of BCK-algebra [1]. Now we follow them and obtain some results such as mentioned in the abstract.

Definition 1. [1, 4] Let $H$ be a non-empty set containing a constant 0 and $\mathcal{P}^{*}(H)$ be the set of all non-empty subset of H . Then $H$ with hyperoperation $\circ: H \times H \rightarrow \mathcal{P}^{*}(H)$, where $(a, b) \mapsto a \circ b$, is called a hyper $K$-algebra (hyper BCK-algebra) if it satisfies $K 1-K 5$ (respectively: HK1-HK4)
K1: $(x \circ z) \circ(y \circ z)<x \circ y, \quad H K 1:(x \circ z) \circ(y \circ z) \ll x \circ y$,
K2: $(x \circ y) \circ z=(x \circ z) \circ y, \quad H K 2:(x \circ y) \circ z=(x \circ z) \circ y$,
K3: $x<x, \quad H K 3: x \circ H \ll x$,
K4: $x<y, y<x$ then $x=y, \quad H K 4: x \ll y, y \ll x$ then $x=y$,
$K 5: 0<x$,

[^7]for all $x, y, z \in H$, where $x<y(x \ll y)$ means $0 \in x \circ y$. Moreover for any $A, B \subseteq H, A<B$ if $\exists a \in A, \exists b \in B$ such that $a<b$ and $A \ll B$ if $\forall a \in A, \exists b \in B$ such that $a \ll b$.

The readers could see some definitions and results about hyper K-algebra and hyper BCK-algebra in $[1,4,5]$.

Definition 2. [6] Let $I$ and $S$ be non-empty subsets of $H$. Then we say that $I$ is $S$-absorbing if $x \in I$ and $y \in S$ then $x \circ y \subseteq I$.

Theorem 1. [6] Let H be a hyper BCK-algebra and I be a hyper BCK-ideal or closed set. Then I is H-absorbing.

Definition 3. [6] A hyper K-algebra $H$ is called ( $P$ )-decomposable if there exists a family $\left\{A_{i}\right\}_{i \in \Omega}$ of subsets of $H$ with $P$-property such that:
(i) $H \neq\left\{A_{i}\right\}$ for all $i \in \Omega$, (ii) $H=\cup_{i \in \Omega} A_{i}$, (iii) $A_{i} \cap A_{j}=\{0\}, i \neq j$.

In this case, we write $H=\oplus_{i \in \Omega} A_{i}(\mathrm{P})$ and say that $\left\{A_{i}\right\}_{i \in \Omega}$ is a (P)decomposition for $H$. If each $A_{i}, i \in \Omega$, is S-absorbing we write $H \stackrel{\underline{\mathrm{~S}}}{\underline{\mathrm{~S}}}$ $\oplus_{i \in \Omega} A_{i}(\mathrm{P})$. Moreover we say that this decomposition is a closed union, in short (P)-CUD, if $\cup_{i \in \Delta} A_{i}$ has P-property for any non-empty subset $\Delta$ of $\Omega$. If there exists a (P)-CUD for $H$, then we say that $H$ is (P)-CUD.

Theorem 2. [6] Let $H \stackrel{H}{=} \oplus_{i \in \Omega} A_{i}$ (a hyper K-ideal). Then $H$ is (hyper $K$-ideal)-CUD.

## 2. Union hyper K-algebras

In this section at first we show that Theorem 3.7 [1] as follows is not true in general, then we give a correct version of it.
Theorem 3.7 of [1]: Let $\left(H_{1}, \mathrm{o}_{2}, 0\right)$ and $\left(H_{2}, \mathrm{o}_{2}, 0\right)$ be hyper $K$-algebras (resp. hyper BCK-algebras) such that $H_{1} \cap H_{2}=\{0\}$ and $H=H_{1} \cup H_{2}$. Then $(H, o, 0)$ is hyper $K$-algebra (resp. hyper BCK-algebras), where the hyperoperation $\circ$ on $H$ is defined as follows:

$$
x \circ y:= \begin{cases}x \circ_{1} y & \text { if } x, y \in H_{1}, \\ x \circ_{2} y & \text { if } x, y \in H_{2}, \\ \{x\} & \text { otherwise, }\end{cases}
$$

for all $x, y \in H$.

Remark 1. This theorem is not true in general. Because we have $0 \circ 0$ is a subset of $H_{1}$ and $H_{2}$. Hence $0 \circ 0$ must be $\{0\}$ in any $H_{i}, i=1,2$, but the authors have not considered this fact, this yields that $\circ$ is not well-defined. To see this, consider the following example.

Example 1. Let $H_{1}=\{0,1,2\}$ and $H_{2}=\{0,3,4\}$. Then $\left(H_{1}, \circ_{1}, 0\right)$ and $\left(H_{2}, \mathrm{o}_{2}, 0\right)$ with hyperoperations $\mathrm{o}_{1}$ and $\mathrm{o}_{2}$ as follows are hyper K-algebras.

| $\circ_{1}$ | 0 | 1 | 2 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0,1\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,1,2\}$ |


| $\circ_{2}$ | 0 | 3 | 4 |
| :---: | :---: | :---: | :---: |
| 0 | $\{0,3,4\}$ | $\{0\}$ | $\{0\}$ |
| 3 | $\{1\}$ | $\{0,3,4\}$ | $\{3\}$ |
| 4 | $\{4\}$ | $\{4\}$ | $\{0,3,4\}$. |

By the definition of hyper operation o for the union of two hyper K-algebras in theorem 3.7 of [1] we have: $(i): 0 \circ 0=\{0,1\}$, because $0 \in H_{1}$. (ii): $0 \circ 0=\{0,3,4\}$, because $0 \in H_{2}$. (iii): $0 \circ 0=\{0\}$, because $0 \in H_{1} \cap H_{2}$. Thus $\circ$ is not well-defined.

Theorem 3.7 [1] is not correct even if we assume $0 \circ 0=\{0\}$. For this, consider the following example.

Example 2. Let $H_{1}=\{0,1,2\}, H_{2}=\{0,3,4\}$ with hyperoperation $\circ_{1}$ and $\mathrm{o}_{2}$ respectively as follows:

| $\circ_{1}$ | 0 | 1 | 2 |
| ---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0,1\}$ | $\{0,1,2\}$ |
| 1 | $\{1\}$ | $\{0,1,2\}$ | $\{0,1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,1,2\}$ |


| $\circ_{2}$ | 0 | 3 | 4 |
| ---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0,3\}$ | $\{0,4\}$ |
| 3 | $\{3\}$ | $\{0,3,4\}$ | $\{0,3,4\}$ |
| 4 | $\{4\}$ | $\{3,4\}$ | $\{0,3,4\}$ |

Then $H_{1}=\left(H, \circ_{1}, 0\right)$ and $H_{2}=\left(H, o_{2}, 0\right)$ are hyper K-algebras and $H_{1} \cap$ $H_{2}=\{0\}$. Thus by considering the hyperoperation $\circ$ of the theorem 3.7 of [1] we have $H=H_{1} \cup H_{2}=\{0,1,2,3,4\}$ with hyperoperation $\circ$ as follows:

| $\circ$ | 0 | 1 | 2 | 3 | 4 |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0,1\}$ | $\{0,1,2\}$ | $\{0,3\}$ | $\{0,4\}$ |
| 1 | $\{1\}$ | $\{0,1,2\}$ | $\{0,1\}$ | $\{1\}$ | $\{1\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,1,2\}$ | $\{2\}$ | $\{2\}$ |
| 3 | $\{3\}$ | $\{3\}$ | $\{3\}$ | $\{0,3,4\}$ | $\{0,3,4\}$ |
| 4 | $\{4\}$ | $\{4\}$ | $\{4\}$ | $\{3,4\}$ | $\{0,3,4\}$ |

But $(H, \circ, 0)$ is not hyper K-algebra. Since $(1 \circ 1) \circ 3=\{0,1,2,3\} \neq$ $(1 \circ 3) \circ 1=\{0,1,2\}$, i.e., K2 does not hold.

Now we give not only a correct version of it, but also we extend it.

Theorem 3. Let $\left(H_{i}, \circ_{i}, 0\right), i \in \Omega$ be a family of hyper $K$-algebras such that $H_{i} \cap H_{j}=\{0\}, i \neq j \in \Omega$ and 0 be a left scalar in each $H_{i}, i \in \Omega$. Then ( $H, \circ, 0$ ) is hyper $K$-algebra where, $H=\cup_{i \in \Omega} H$ and $\circ$ on $H$ is defined as follows:

$$
x \circ y:= \begin{cases}x \circ_{i} y & \text { if } x, y \in H_{i}, \\ \{x\} & \text { if } x \in H_{i}, y \notin H_{i} .\end{cases}
$$

Proof. Since 0 is a left scalar in $H_{i}$ for all $i \in \Omega$, then $0 \circ 0=\{0\}$ and $\circ$ is well-defined. Now we prove $H$ satisfies K1-K5. If $x, y, z \in H_{i}$ for some $i \in \Omega$ then, by hypothesis, $H$ satisfies $K 1-K 5$, otherwise we consider the following cases:
(I) $x \in H_{i}$ and $y, z \notin H_{i}$,
(II) $x, y \in H_{i}$ and $z \notin H_{i}$,
(III) $x, z \in H_{i}$ and $y \notin H_{i}$.

K1: $(x \circ z) \circ(y \circ z)<x \circ y, \forall x, y, z \in H$.
If (I) holds and $0 \notin y \circ z$, then $(x \circ z) \circ(y \circ z)=x \circ(y \circ z)=\{x\}<x \circ y=\{x\}$. Otherwise $(x \circ z) \circ(y \circ z)=x \circ(y \circ z)=x \circ 0<x \circ y=\{x\}$, since $x \in x \circ 0$. If (II) holds, then $(x \circ z) \circ(y \circ z)=x \circ_{i} y<x \circ i y$. If (III) holds, by considering $x \circ z \subseteq H_{i}$ and 0 is a left scalar, then $(x \circ z) \circ(y \circ z)=(x \circ z) \circ y=x \circ z<$ $x \circ y=\{x\}$.

K2: $(x \circ y) \circ z=x \circ y=(x \circ z) \circ y, \forall x, y, z \in H$.
If (I) holds, then $(x \circ y) \circ z=x \circ z=\{x\}=(x \circ z) \circ y$. If (II) holds, then $(x \circ y) \circ z=x \circ y=(x \circ z) \circ y$, even if $0 \in x \circ y$. Since 0 is a left scalar and $x \circ y \subseteq H_{i}$, so we have $(x \circ y) \circ z=x \circ i y$ and K2 holds. If (III) holds, then as proof (II) we have $(x \circ y) \circ z=x \circ_{i} z=(x \circ z) \circ y$, and K2 holds.

The proof of K3 and K5 are straightforward.
K4: If $x<y$ and $y<x$, then $x=y$. Let $x, y \in H$ be such that $x<y$ and $y<x$. We consider two cases (i): $x \in H_{i}, y \notin H_{i}$ and (ii): $x, y \in H_{i}$. If (i) holds, then $0 \in x \circ y=\{x\}$ and $0 \in y \circ x=\{y\}$, hence $x=y=0$. If (ii) holds, then $H$ satisfies K4. Therefore $(H, \circ, 0)$ is a hyper K-algebra.

Definition 4. Let $\left(H_{i}, \circ_{i}, 0\right), i \in \Omega$ be hyper K-algebras such that $H_{i} \cap$ $H_{j}=\{0\}, i \neq j \in \Omega$ and 0 be a left scalar in each $H_{i}, i \in \Omega$. Then the hyper K-algebra $(H, \circ, 0)$ which has been defined in Theorem 3 is called the union of a family $\left\{H_{i}: i \in \Omega\right\}$ of hyper K-algebras and it is denoted by $H=\oplus_{i \in \Omega} H_{i}$ (hyper K-algebra).

Now we consider some properties of $H_{i}$ 's, $i \in \Omega$ in which they can be extended to $H=\oplus_{i \in \Omega} H_{i}$ (hyper K-algebra).

Theorem 4. Let $H=\oplus_{i \in \Omega} H_{i}$ (hyper K-algebra). Then
(i) whenever 0 is a right scalar, the hyper $K$-algebra $H$ is positive implicative if and only if each $H_{i}, i \in \Omega$ is positive implicative;
(ii) whenever 0 is a right scalar, the hyper $K$-algebra $H$ is strong implicative if and only if each $H_{i}, i \in \Omega$ is strong implicative;
(iii) the hyper K-algebra $H$ is weak implicative (implicative) if and only if each $H_{i}, i \in \Omega$ is weak implicative (implicative).

Proof. Since the proof of $(\Rightarrow)$ is clear, we prove only $(\Leftarrow)$.
(i) Let each $H_{i}, i \in \Omega$ satisfies the identity:

$$
\begin{equation*}
(x \circ y) \circ z=(x \circ z) \circ(y \circ z) \tag{1}
\end{equation*}
$$

We have to show that $H$ satisfies (1) for all $x, y, z \in H$. For briefly, we denote $(x \circ y) \circ z$ by A and $(x \circ z) \circ(y \circ z)$ by B and proceed the proof by following cases.
Case 1: If $x, y, z \in H_{i}$ for some $i \in \Omega$, then the proof is clear.
Case 2: If $x, y \in H_{i}$ and $z \in H_{j}$ where $i \neq j \in \Omega$, then from the definition of $\circ$ and the fact that 0 is a left scalar in each $H_{i}$, we have $A=x \circ_{i} y=B$.
Case 3: If $x, z \in H_{i}$ and $y \in H_{j}$ where $i \neq j$, then by K2 and Case 2 the proof is obvious.
Case 4: If $x \in H_{i}$ and $y, z \in H_{j}$ where $i \neq j$, then $A=\{x\}$. Since $y \circ z \subseteq H_{j}$ and 0 is a right scalar, so we have $\mathrm{B}=x \circ(y \circ z)=\{x\}$, hence $\mathrm{A}=\mathrm{B}$.
Case 5: If $x \in H_{i}, y \in H_{j}$ and $z \in H_{k}$ where $i \neq j, i \neq k, j \neq k$ and $i, j, k \in \Omega$, then $A=\{x\}=B$, which completes the proof of $(i)$.
(ii) We know $H$ is strong implicative if $x \circ 0 \subseteq x \circ(y \circ x)$. Let $x, y \in H$. For $x, y \in H_{i}, i \in \Omega$, the proof is obvious. For $x \in H_{i}, y \in H_{j}$, where $i \neq$ $j \in \Omega$, the fact that 0 is a right scalar implies $x \circ 0=\{x\} \subseteq x \circ(y \circ x)=\{x\}$, which completes the proof.
(iii) If $H$ is weak implicative (implicative) then $x<x \circ$ ( $y \circ x$ ) (resp. $x \in x \circ(y \circ x))$. Thus the case $x, y \in H_{i}, i \in \Omega$, is obvious. In the case $x \in H_{i}, y \in H_{j}, i \neq j \in \Omega$, we have $x \circ(y \circ x)=\{x\}$. Hence $x<x \circ(y \circ x)$ $(\operatorname{resp} . x \in x \circ(y \circ x))$.

Theorem 5. Let $H=\oplus_{i \in \Omega} H_{i}$ (hyper K-algebra). Then $H \stackrel{H}{=} \oplus_{i \in \Omega} H_{i}$ (hyper $K$-ideal), moreover $H$ is (hyper $K$-ideal)-CUD.

Proof. It is sufficient to prove that each $H_{i}, i \in \Omega$, is an H-absorbing hyper K-ideal in $H$. By Theorem 2, $H$ is (hyper K-ideal)-CUD. Let $x \neq 0, x \circ y<$
$H_{i}$ and $y \in H_{i}$. If $x \notin H_{i}$, then we have $0 \in(x \circ y) \circ t=\{x\}$, for some $t \in H_{i}$, hence $x=0$ which is a contradiction. So, $x \in H_{i}$ and $H_{i}$ is a hyper K-ideal of $H$. Since $H_{i} \cap H_{j}=\{0\}, i \neq j$, and $H=\cup H_{i \in \Omega}$, by Definition 3 we conclude that $H=\oplus_{i \in \Omega} H_{i}$ (hyper K-ideal). Also $H_{i}, i \in \Omega$ is H -absorbing. Indeed, for $x, y \in H_{i}$ we have $x \circ y=x \circ_{i} y \subseteq H_{i}$. For $x \in H_{i}, y \notin H_{i}$, we have $x \circ y=\{x\} \subset H_{i}$. Hence each $H_{i}$ is H -absorbing.

Corollary 1. Let $H=\oplus_{i \in \Omega} H_{i}$ (hyper K-algebra). Then $H \stackrel{H}{\underline{\mathrm{H}}} \oplus_{i \in \Omega} H_{i}$ (closed set), moreover $H$ is (closed set)-CUD

Proof. Since a hyper K-ideal is closed set, the proof follows from Theorem 5.

Now we show that there exists a (hyper K-ideal)-decomposable hyper K-algebra such that it is not a union of any family of hyper K-algebras. Hence the converse of Theorem 5, is not true in general. In the next section we show that if $H$ is (hyper BCK-ideal)-decomposable, then it is a union of a family of hyper BCK-algebras.

Example 3. Let $H=\{0,1,2,3\}$ and consider the following table:

| $\circ$ | 0 | 1 | 2 | 3 |
| :---: | :---: | :---: | :---: | :---: |
| 0 | $\{0\}$ | $\{0\}$ | $\{0\}$ | $\{0\}$ |
| 1 | $\{1\}$ | $\{0,1\}$ | $\{0,1\}$ | $\{1,2\}$ |
| 2 | $\{2\}$ | $\{2\}$ | $\{0,2\}$ | $\{2\}$ |
| 3 | $\{3\}$ | $\{3\}$ | $\{3\}$ | $\{0,3\}$ |

Then $(H, \circ, 0)$ is a hyper K-algebra and $H=\{0,1,2\} \oplus\{0,3\}$. It is clear that $\{0,1,2\}$ and $\{0,3\}$ are hyper K -subalgebras of $H$, but $H$ is not the union of $H_{1}=\{0,1,2\}$ and $H_{2}=\{0,3\}$ because $1 \circ 3=\{1,2\} \neq\{1\}$.

## 3. Union hyper BCK-algebras

In this section we show that a hyper BCK-algebra is (closed set)-decomposable if and only if it is the union of a family of hyper BCK-algebras.

Lemma 1. Let $(H, \circ, 0)$ be a hyper BCK-algebra. If $0 \in(x \circ u) \circ 0$ where $x, u \in H$, then $x \ll u$.

Proof. Suppose $0 \in(x \circ u) \circ 0$, then there is an element $t \in x \circ u$ in which $0 \in t \circ 0$, that is, $t \ll 0$. Since $0 \ll t$ and HK4 holds, we conclude that $t=0$. Since $t \in x \circ u$, thus $0 \in x \circ u$ or $x \ll u$.

Theorem 6. In any hyper BCK-algebra 0 is a scalar.
Proof. In any hyper BCK-algebra we have $0 \circ x=\{0\}$, i.e., 0 is a left scalar. Now we show that 0 is a right scalar, i.e., $x \circ 0=\{x\}$. We know $x \in x \circ 0$. Now let $u \in x \circ 0$. Then we show that $u=x$. From $0 \in u \circ u$ and HK2 we get that $0 \in(x \circ 0) \circ u=(x \circ u) \circ 0$. Hence by Lemma $1, x \ll u$. On the other hand we have $x \circ 0 \ll x$, so $u \ll x$. By considering HK4, from $u \ll x$ and $x \ll u$ we conclude that $u=x$. Thus $x \circ 0=\{x\}$.

Theorem 7. Let $\left(H_{i}, \circ_{i}, 0\right), i \in \Omega$ be hyper BCK-algebras such that $H_{i} \cap H_{j}$ $=\{0\}, i \neq j \in \Omega$. Then $(H, \circ, 0)$, where $H=\cup_{i \in \Omega} H$ and

$$
x \circ y:=\left\{\begin{array}{lll}
x \circ_{i} y & \text { for } x, y \in H_{i} \\
\{x\} & \text { for } & x \in H_{i}, y \notin H_{i},
\end{array}\right.
$$

is a hyper BCK-algebra. We denote it by $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-algebra).
Proof. By Lemma 6, the element 0 is a scalar in any hyper BCK-algebra, hence the proof follows from Theorem 3.

Theorem 8. Let $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-algebra). Then each $H_{i}, i \in \Omega$ is weak implicative (implicative, strong implicative) if and only if $H$ is weak implicative (implicative, strong implicative).

Proof. Since 0 is a scalar, the proof follows from Theorem 4.
Lemma 2. Any closed subset of a hyper BCK-algebra is a hyper BCKsubalgebra.

Theorem 9. (Main Theorem) Let $(H, \circ, 0)$ be a hyper BCK-algebra. Then $H=\oplus_{i \in \Omega} A_{i}$ (closed set) if and only if $H=\oplus_{i \in \Omega} A_{i}$ (hyper BCKalgebra).

Proof. $(\Rightarrow)$ Let $H=\oplus_{i \in \Omega} A_{i}$ (closed set). Then by Lemma $2, A_{i}$ is a hyper BCK-subalgebra for any $i \in \Omega$. Now, suppose $0 \neq x \in A_{i}$ and $y \notin A_{i}$. In view of Theorem 7, we must show $x \circ y=\{x\}$. We assume that $y \in A_{j}$ where $i \neq j \in \Omega$. If $0 \in x \circ y$, i.e., $x \ll y$ then $x \in A_{j}$, because $y \in A_{j}$ and $A_{j}$ is a closed set. This is a contradiction to $A_{i} \cap A_{j}=\{0\}$. So, let $0 \notin x \circ y$. We know $x \circ y \ll x$. Let $u \in x \circ y$. Then $u \ll x$, and $0 \in(x \circ y) \circ u=(x \circ u) \circ y$. Therefore $0 \in(x \circ u) \circ y$. Hence there exists $t \in x \circ u$ in which, $0 \in t \circ y$, i.e., $t \ll y$. Since $y \in A_{j}$ and $A_{j}$ is a closed set we get that $t \in A_{j}$. By Theorem 1, as $A_{i}$ is H -absorbing, we have $t \in x \circ u \subseteq A_{i}$ and consequently
$t \in A_{i} \cap A_{j}$. This implies that $t=0$ and hence $0 \in x \circ u$, i.e., $x \ll u$. On the other hand we had $u \ll x$, these imply that $x=u$. Therefore $x \circ y=\{x\}$. $(\Leftarrow)$ The proof follows from Corollary 1.

Corollary 2. Let $(H, \circ, 0)$ be a hyper BCK-algebra. Then $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-algebra) if and only if $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-ideal).

Corollary 3. Let $(H, \circ, 0)$ be a hyper BCK-algebra. Then $H=\oplus_{i \in \Omega} H_{i}$ (hyper BCK-ideal) if and only if $H=\oplus_{i \in \Omega} H_{i}$ (closed set).

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# On decomposition of commutative Moufang groupoids 

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#### Abstract

We prove that every commutative Moufang groupoid is a semilattice of Archimedean subgroupoids.


It is well-known that the multiplicative groupoid of an alternative/Jordan algebra satisfies Moufang identities [1, 4]. Therefore it seems interesting to study the structure of such groupoids. In this note we apply to Moufang groupoids an approach which is widespread in semigroup theory decomposition into a semilattice of subsemigroups [3].

We shall call a groupoid with the identity

$$
\begin{equation*}
(x y)(z x)=(x(y z)) x \tag{1}
\end{equation*}
$$

## a Moufang groupoid. Everywhere in this article $M$ denotes a commutative Moufang groupoid.

Theorem 1. If $M$ consists of idempotents, then it is a semilattice.
Proof. Under assumption of the theorem it follows from (1) for $y=z$

$$
\begin{equation*}
(x y) x=x y \tag{2}
\end{equation*}
$$

Applying (2) to the right part of (1), we get:

$$
\begin{equation*}
x(y z)=(x y)(x z) \tag{3}
\end{equation*}
$$

Now define a binary relation $\leqslant$ on $M$ :

$$
a \leqslant b \Longleftrightarrow a b=a
$$

and show that it is a partial order.
Indeed, the reflexivity follows from idempotentness, the antisymmetry follows from commutativity. Let $a \leqslant b \leqslant c$. Then

$$
a c=(a b) c=(a c)(b c)=(a c) b=(a b)(b c)=a b=a,
$$

i.e., $a \leqslant c$.

Further, $a b$ is a greatest lower bound for the pair $\{a, b\}$. Really, $a b \leqslant a$, $a b \leqslant b$ by (2). Suppose that $x \leqslant a, x \leqslant b$. Then $(a b) x=(a x)(b x)=x \cdot x=$ $x$, i.e., $x \leqslant a b$.

Lemma 2. $M$ is a groupoid with associative powers.
Proof. For $a \in M$ we shall denote by $a^{(n)}$ an arbitrary term of the length $n \geq 1$, all letters of which are $a$. If all such terms coincide in $M$, we denote them by $a^{n}$.

We use the induction on length of the term. Let $a^{(k)}=a^{k}$ for any $k<n$ (for $k=3$ this follows from commutativity). Consider some term $a^{(n)}$. It can be written in the form $a^{(n)}=a^{(k)} a^{(l)}$, where $k, l \geqslant 1$ and $k+l=n$; in view of commutativity one can assume that $k \leqslant l$.

Suppose that $k \geqslant 2$. Then under hypothesis of the induction

$$
a^{(n)}=a^{k} a^{l}=\left(a a^{k-1}\right)\left(a a^{l-1}\right)=\left(a\left(a^{k-1} a^{l-1}\right)\right) a=\left(a a^{n-2}\right) a=a a^{n-1} .
$$

Hence all terms of the form $a^{(n)}$ are equal.
We denote by $L_{a}$ the left translation corresponding to an element $a$ : $L_{a} b=a b$. From (1) we have:

$$
(x y)^{2}=L_{x}^{2} y^{2} .
$$

We generalize this identity:
Lemma 3. $(a b)^{2^{n}}=L_{a}^{2^{n}} b^{2^{n}}$ for any $a, b \in M, n \geqslant 0$ (here powers are defined correctly in view of Lemma 2).

Proof. Assume that for $n$ the statement is faithful and prove it for $n+1$ :
$(a b)^{2^{n+1}}=\left[(a b)^{2}\right]^{2^{n}}=\left[a\left(a b^{2}\right)\right]^{2^{n}}=L_{a}^{2^{n}}\left(a b^{2}\right)^{2^{n}}=L_{a}^{2^{n}} L_{a}^{2^{n}} b^{2^{n+1}}=L_{a}^{2^{n+1}} b^{2^{n+1}}$

Corollary 4. $\left(L_{a_{1}} \ldots L_{a_{k-1}} a_{k}\right)^{2^{n}}=L_{a_{1}}^{2^{n}} \ldots L_{a_{k-1}}^{2^{n}} a_{k}^{2^{n}}$.

Further we shall need one more equality for translations:
Lemma 5. $L_{a}^{2 n} L_{b}=L_{L_{a}^{n} b} L_{a}^{n}$ for any $a, b \in M, n \geqslant 1$.
Proof. For $n=1$ this statement coincides with (1). The general case is obtained by induction on $n$.

Let $I_{a}$ be denoted the principal ideal, generated by $a \in M$. It is clear that each element from $I_{a}$ can be written in the form $L_{x_{1}} \ldots L_{x_{k-1}} L_{x_{k}} a$.

Define relations $\rho$ and $\sigma$ :

$$
\begin{gather*}
a \rho b \Longleftrightarrow \exists n \geqslant 1 \quad a^{n} \in I_{b},  \tag{4}\\
a \sigma b \tag{5}
\end{gather*}
$$

Lemma 6. $\sigma$ is a congruence.
Proof. Reflexivity and symmetry are obvious, it is enough to check transitivity and stability of $\rho$. Note that one can assume in the definition of $\rho$ that $n$ is the power of the two.

Let $a \rho b, b \rho c$, i.e.,

$$
a^{2^{m}}=L_{x_{1}} \ldots L_{x_{k}} b, \quad b^{2^{n}}=L_{y_{1}} \ldots L_{y_{l}} c
$$

By Corollary 4

$$
a^{2^{m+n}}=L_{x_{1}}^{2^{n}} \ldots L_{x_{k}}^{2^{n}} 2^{2 n}=L_{x_{1}}^{2^{n}} \ldots L_{x_{k}}^{2^{n}} L_{y_{1}} \ldots L_{y_{l}} c \in I_{c},
$$

so $\rho$ is transitive.
Now let $a \rho b$, i.e., $a^{2^{n}}=L_{x_{1}} \ldots L_{x_{k}} b$, and $c \in M$.

1) Suppose that $k \leqslant n$. Then using several times Lemma 5, we get for some $u_{1}, \ldots, u_{k} \in M$ :

$$
(c a)^{2^{n}}=L_{c}^{2^{n}} a^{2^{n}}=L_{c}^{2^{n}} L_{x_{1}} \ldots L_{x_{k}} b=L_{u_{1}} \ldots L_{u_{k}} L_{c}^{2^{n-k}} b \in I_{c b} .
$$

2) Let $k>n$. Then $a^{2^{n+k+1}}=L_{y} L_{x_{1}} \ldots L_{x_{k}} b$, where $y=a^{2^{n+k+1}-2^{n}}$. Since $k+1<n+k+1$, we get the case 1). Consequently, capcb.

Lemma 7. $M / \sigma$ is a semilattice.
Proof. Obviously, $a \sigma a^{2}$ for any $a \in M$. So $M / \sigma$ is an idempotent groupoid. By Theorem 1 it is a semilattice.

Now let us to consider the structure of $\sigma$-classes (of course, they are subgroupoids).

Like to theory of semigroups, we call a groupoid $M$ Archimedean if $a \sigma b$ for any $a, b \in M$, where $\sigma$ is defined by the conditions (4) and (5). It is clear that an Archimedean groupoid is indecomposable into a semilattice of subgroupoids.

Lemma 8. Let $\sigma$ be a congruence on $M$, defined by conditions (4) and (5). Then each $\sigma$-class is Archimedean.

Proof. Let $N$ is a $\sigma$-class, $a, b \in N$. Then

$$
\begin{equation*}
a^{n}=L_{x_{1}} \ldots L_{x_{k}} b \tag{6}
\end{equation*}
$$

for some $n>0, x_{1}, \ldots, x_{k} \in M$. We need to prove that in the equality (6) elements $x_{1}, \ldots, x_{k}$ can be chosen from $N$.

From (6) and Lemma 5 we have:

$$
a^{n+2^{k}}=L_{a}^{2^{k}} L_{x_{1}} \ldots L_{x_{k}} b=L_{L_{a}^{2 k-1} x_{1}} L_{L_{a}^{2 k-2} x_{2}} \ldots L_{L_{a} x_{k}} b .
$$

Show that for any $i \leqslant k$ the element $y_{i}=L_{a}^{2^{k-i}} x_{i}$ is contained in $N$. Indeed, since $y_{i}=a\left(L_{a}^{2^{k-i}-1} x_{i}\right)$ then $y_{i} \rho a$. On the other hand,

$$
a^{n+2^{k}}=L_{y_{1}} \ldots L_{y_{k}} b=L_{y_{1}} \ldots L_{y_{i-1}}\left[\left(L_{y_{i+1}} \ldots L_{y_{k}} b\right) y_{i}\right],
$$

whence $a \rho y_{i}$. Thereby, $a \sigma y_{i}$, i.e., $y_{i} \in N$.
The final result:
Theorem 9. A commutative Moufang groupoid is a semilattice of Archimedean groupoids.

Example. Let a finite semigroup $S$ satisfy the identity $a b=a$ ( $a$ left zero semigroup), $F$ be a field, char $F \neq 2, A=F S$ be the semigroup algebra. $A$ is a Jordan algebra with respect to the operation $x * y=\frac{1}{2}(x y+y x)$. Denote by $A^{*}$ its multiplicative groupoid (as is well-known it is Moufang and commutative [4]).

The operation in $A^{*}$ can be written as follows. For $x=\sum_{a \in S} \alpha_{a} a \in A^{*}$, $\alpha_{a} \in F$, denote $|x|=\sum_{a \in S} \alpha_{a}$. Then

$$
x * y=\frac{1}{2}(|x| y+|y| x)
$$

From here $x^{* n}=|x|^{n-1} x$ and $|x y|=|x||y|$. In particular, $x \in \operatorname{Rad} A^{*}$ iff $|x|=0$.

Evidently, all elements from $\operatorname{Rad} A^{*}$ constitute one $\sigma$-class. On the other hand, if $x, y \notin \operatorname{Rad} A^{*}$ then they divide one another. To make sure that, it is enough to put

$$
t=\frac{1}{|x|^{2}}(2|x| y-|y| x) ;
$$

then $y=x * t$. Thus $A^{*}=\operatorname{Rad} A^{*} \cup\left(A^{*} \backslash \operatorname{Rad} A^{*}\right)$ is the decomposition of $A^{*}$ into Archimedean components.

Finally we discuss some problems which arise here.

1. For loops the identity (1) (central Moufang identity) is equal to each of ones $x(y(x z))=((x y) x) z$ and $((z x) y) x=z(x(y x))$ (left and right Moufang identities). This is valid for multiplicative groupoids of Jordan algebras as well, but not in the general case. So we can consider left and right Moufang (commutative) groupoids. Are there similar decompositions for them?
2. Is there Archimedean decomposition in noncommutative situation? This is the case for semigroups [2].
3. What can one say about the structure of an Archimedean component? For instance, can it contain more than one idempotent (cf. [3], Ex.4.3.2)?

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# Greedy quasigroups 

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#### Abstract

The paper investigates the quasigroup $Q_{s}$ constructed on the well-ordered set of natural numbers by placing a number $s$ known as the seed in the top left-hand corner of the body of the multiplication table, and then completing the Latin square using the greedy algorithm that chooses the least possible entry at each stage. The initial motivation comes from the theory of combinatorial games, where $Q_{0}$ gives the usual nim sum, while $Q_{1}$ gives the corresponding sums for positions in misère nim. The multiplication groups of these quasigroups are analyzed. The alternating group of the natural numbers is a subgroup of the multiplication groups. It is shown that these so-called greedy quasigroups $Q_{s}$ are mutually non-isomorphic. The quasigroup $Q_{1}$ is subdirectly irreducible. For $s>1$, the greedy quasigroups $Q_{s}$ are simple, and for $s>2$ they are rigid, possessing no non-trivial automorphisms. Indeed in this case the endomorphism monoid contains just the identity and a single constant. The subquasigroup structures of the $Q_{s}$ are also determined. While $Q_{0}, Q_{1}$ have uncountably many subquasigroups, and $Q_{2}$ has just one proper, non-trivial subquasigroup, $Q_{s}$ has none for $s>2$.


## 1. Introduction

In this paper, quasigroups motivated by combinatorial games, nim in particular, are examined. They form a countably infinite family of infinite quasigroups with some curious properties. The underlying set $Q$ of the quasigroups is taken to be the well-ordered set of natural numbers including 0. A quasigroup is constructed by filling in the multiplication table in a greedy fashion with the rows and columns labelled by the elements of $Q$ in their natural order. For each proper subset $S$, of $Q$, define the minimal

[^8]excluded number mex $S$ of $S$ to be the least element of the (non-empty) complement of $S$ in $Q$. (This element is uniquely defined by the well-ordering principle.) Fix a natural number $s$, known as the seed. Define
\[

$$
\begin{equation*}
0 \cdot 0:=s . \tag{1}
\end{equation*}
$$

\]

One may then use the following greedy algorithm to define the remaining products of natural numbers $l$ and $m$ inductively:

$$
\begin{equation*}
l \cdot m:=\operatorname{mex}(\{i \cdot m \mid i<l\} \cup\{l \cdot j \mid j<m\}) . \tag{2}
\end{equation*}
$$

The algorithm guarantees that the body of the multiplication table will be a (infinite) Latin square, and therefore that $(Q, \cdot)$ becomes a quasigroup $Q_{s}$, known as the greedy quasigroup seeded by $s$. As an illustration, the following table

|  | 0 | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 | 9 | 10 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 5 | 0 | 1 | 2 | 3 | 4 | 6 | 7 | 8 | 9 | 10 |
| 1 | 0 | 1 | 2 | 3 | 4 | 5 | 7 | 6 | 9 | 8 | 11 |
| 2 | 1 | 2 | 0 | 4 | 5 | 3 | 8 | 9 | 6 | 7 | 12 |
| 3 | 2 | 3 | 4 | 0 | 1 | 6 | 5 | 8 | 7 | 10 | 9 |
| 4 | 3 | 4 | 5 | 1 | 0 | 2 | 9 | 10 | 11 | 6 | 7 |
| 5 | 4 | 5 | 3 | 6 | 2 | 0 | 1 | 11 | 10 | 12 | 8 |
| 6 | 6 | 7 | 8 | 5 | 9 | 1 | 0 | 2 | 3 | 4 | 13 |
| 7 | 7 | 6 | 9 | 8 | 10 | 11 | 2 | 0 | 1 | 3 | 4 |
| 8 | 8 | 9 | 6 | 7 | 11 | 10 | 3 | 1 | 0 | 2 | 5 |
| 9 | 9 | 8 | 7 | 10 | 6 | 12 | 4 | 3 | 2 | 0 | 1 |
| 10 | 10 | 11 | 12 | 9 | 7 | 8 | 13 | 4 | 5 | 1 | 0 |

Table 1. Part of the multiplication table of $Q_{5}$.
gives the first few entries of the multiplication table of $Q_{5}$.
Seeding with 0 , one obtains $Q_{0}$ as a countable elementary abelian 2group. In the theory of combinatorial games, the multiplication of $Q_{0}$ is known as nim sum [5]. Greedy quasigroups will be seen as a generalization of nim. Each position $X$ in the game of nim is assigned a natural number value $x$, and the nim sum $x \oplus y$ denotes the value of the nim position $X+Y$ obtained by juxtaposing $X$ with a second position $Y$ of value $y$. Seeding with 1, the quasigroup $Q_{1}$ gives a comparable description of the juxtaposition of positions in the game of misère nim - nim played to lose. Section
discusses elementary properties of the greedy quasigroups: commutativity, associativity, total symmetry, and the existence of identity, idempotent, and nilpotent elements. The next two sections (which may be skipped at first reading) comprise a number of technical lemmas about the multiplication by 2 and 3 in $Q_{s}$ for $s>0$. These lemmas drive the theorems regarding the multiplication groups in Section. Section examines the subquasigroup structure of the greedy quasigroups. It transpires that while $Q_{0}, Q_{1}$ have uncountably many subquasigroups, $Q_{2}$ has just one proper, non-trivial subquasigroup, and $Q_{s}$ has none for $s>2$ (Theorem 6.2). It is also shown that for $s>2$, the quasigroup $Q_{s}$ is simple. The congruences of $Q_{0}$ correspond directly to its subgroups, essentially forming a projective geometry of countable dimension over the 2-element field. For $s>1$, the greedy quasigroups $Q_{s}$ are shown to be simple in Theorem 6.3. Section considers homomorphisms between greedy quasigroups. It is shown that the greedy quasigroups are mutually non-isomorphic (Theorem 7.1), and indeed that for distinct positive seeds $s, t$, the only homomorphism from $Q_{s}$ to $Q_{t}$ is the constant map taking the value 1 (Theorem 7.11). Finally, Theorem 7.12 shows that for $s>2$, the only endomorphisms of $Q_{s}$ are the constant and the identity. In particular, $Q_{s}$ is rigid in the sense of having a trivial automorphism group. It may be worth noting that the properties of the greedy quasigroups $Q_{s}$ for $s>2$, namely simplicity, rigidity, and lack of proper, non-trivial subalgebras, are reminiscent of the Foster-Pixley characterization of (necessarily finite) primal algebras [7]. The paper concludes with a brief characterization of greedy quasigroups in terms of combinatorial game theory. For algebraic concepts and conventions that are not otherwise explained here, especially involving quasigroups, readers are referred to [8]. Note that mappings are usually placed to the right of their arguments, allowing composition in natural order, and minimizing the number of brackets that otherwise proliferate in the study of non-associative structures such as quasigroups.

## 2. Elementary properties

Recall that a quasigroup $(Q, \cdot, /, \backslash)$ is said to be commutative or associative respectively if its multiplication • is commutative or associative.

Proposition 2.1. For each seed $s$, the quasigroup $Q_{s}$ is commutative.
Proof. By induction, using (2):

$$
\begin{aligned}
l \cdot m & =\operatorname{mex}(\{i \cdot m \mid i<l\} \cup\{l \cdot j \mid j<m\}) \\
& =\operatorname{mex}(\{m \cdot i \mid i<l\} \cup\{j \cdot l \mid j<m\}) \\
& =\operatorname{mex}(\{i \cdot l \mid i<m\} \cup\{m \cdot j \mid j<l\})=m \cdot l .
\end{aligned}
$$

(The induction hypothesis is used for the second equality.)
Proposition 2.2. Suppose $s>0$.

1. $\forall 0<x \leqslant s, 0 \cdot x=x \cdot 0=x-1$.
2. $\forall x>s, 0 \cdot x=x \cdot 0=x$.
3. $\forall 0 \leqslant x \leqslant s, 1 \cdot x=x \cdot 1=x$.

Proof. (1) Since $0 \cdot 0=s, 1 \cdot 0=0$, and applying (2) to each successive term, one has $x \cdot 0=\operatorname{mex}\{s, 0,1, \ldots,(x-1) \cdot 0=(x-2)\}=x-1$.
(2) For $x=s+1$, (2) gives $0 \cdot x=\operatorname{mex}\{s, 0,1, \ldots, s-1\}=s+1$. Then $0 \cdot x=x$ for $x>s$ by induction.
(3) Note $0 \cdot 1=0$. Then for $x \leqslant s$, induction yields

$$
x \cdot 1=\operatorname{mex}\{0,1, \ldots, x-1,0 \cdot x=x-1\}=x .
$$

Corollary 2.3. For $s>0$, the quasigroup $Q_{s}$ is not associative.
Proof. $(0 \cdot 0) \cdot(s+1)=s \cdot(s+1) \neq 0 \cdot(s+1)=0 \cdot(0 \cdot(s+1))$.
Definition 2.4. The hub of a greedy quasigroup $Q_{s}$ is defined to be the subset $H_{s}=\{0, \ldots, s\}$.

In Table 1 , the hub $H_{5}$ is marked off by separating lines.
Remark 2.5. The element 0 is the identity element of the group $Q_{0}$. For $s>0$, the quasigroup $Q_{s}$ does not have a universal identity element. However, the later parts of Proposition 2.2 may be interpreted as saying that 1 is an identity for the hub, while 0 is an identity outside the hub. In particular, 1 is the only idempotent element of $Q_{s}$, i.e., the only element $x$ forming a singleton subquasigroup $\{x\}$.

A quasigroup $(Q, \cdot, /, \backslash)$ is said to be totally symmetric if its three binary operations agree, i.e., if the implication

$$
\begin{equation*}
x_{1} \cdot x_{2}=x_{3} \Rightarrow x_{1 \pi} \cdot x_{2 \pi}=x_{3 \pi} \tag{3}
\end{equation*}
$$

holds for all permutations $\pi$ of the index set $\{1,2,3\}$. (Commutativity means that (3) holds for $\pi=$ (12).) Note that $Q_{0}$, like any elementary abelian 2 -group, is totally symmetric. Now outside the hub, the multiplication on $Q_{1}$ is constructed exactly as in $Q_{0}$. Furthermore, the hub of $Q_{1}$ is totally symmetric, being isomorphic to the subgroup $\{0,1\}$ of $Q_{0}$. Thus $Q_{1}$ is also totally symmetric.

Lemma 2.6. Suppose $s>0$. For $x>s$,

$$
x \cdot 1= \begin{cases}x+1, & x-s \equiv_{2} 1 \\ x-1, & x-s \equiv_{2} 0 .\end{cases}
$$

Proof. As an induction basis, note:

$$
\begin{aligned}
& (s+1) \cdot 1=\operatorname{mex}\{0,1, \ldots, s,(s+1) \cdot 0=s+1\}=s+2 \\
& (s+2) \cdot 1=\operatorname{mex}\{0,1, \ldots, s, s+2,(s+2) \cdot 0\}=s+1
\end{aligned}
$$

Consider $x>s$. By induction, for $x-s \equiv_{2} 1$,

$$
x \cdot 1=\operatorname{mex}\{0,1, \ldots, x-1, x \cdot 0\}=x+1,
$$

and for $x-s \equiv_{2} 0$,

$$
x \cdot 1=\operatorname{mex}\{0,1,2, \ldots, x-3+1, x-2-1, x-1+1, x \cdot 0\}=x-1 .
$$

Recall that in any quasigroup ( $Q, \cdot, /, \backslash$ ), the square $x^{2}$ of an element $x$ is $x \cdot x$. An element of a greedy quasigroup is described as nilpotent if its square is 0 . All but at most two elements of a greedy quasigroup are nilpotent, and 0 is the only square of infinitely many elements.

Theorem 2.7. For $x>1$ in any greedy quasigroup, $x^{2}=0$.
Proof. The result is immediate in $Q_{0}$, so suppose $s>0$. Recall $0 \cdot 1=0=$ $1 \cdot 0$. Thus the first place 0 can appear in the 2 -column of the Latin square is the 2 -row, so it must appear there. Then the first place 0 can and must appear in the 3 -column is the 3 -row. Fill in the first $n$ columns (labelled $0, \ldots, n-1$ ) by induction. The first place 0 can appear in the $n$-column is in the $n$-row. Thus by induction $n \cdot n=0$ for all $n>1$.

Corollary 2.8. Consider the greedy quasigroup $Q_{s}$.

1. If $s=1$, then $0^{2}=1^{2}=1$.
2. For $s \neq 1$, the element 0 is the only square of more than one element.

## 3. Multiplication by 2

Throughout the next two technical sections, which may be skipped at first reading, assume $s>0$. (Later, it will be implicitly necessary to assume that $s$ is "sufficiently large.") Consider the inductive construction of the Latin square that forms the body of the multiplication table of $Q_{s}$. There is a critical dependence on the congruence class of the seed to certain moduli. A column is said to be complete at entry $n$ if its first $n+1$ elements are precisely the numbers $0,1, \ldots, n$. The proofs are by induction and can be done by hand in a similar fashion to those above.

Lemma 3.1. For $x<s$,

$$
x \cdot 2= \begin{cases}x+1, & x \equiv_{3} 0,1 ; \\ x-2, & x \equiv_{3} 2 .\end{cases}
$$

The post-hub behavior of the 2-column depends on the congruence class of the seed modulo 3. We consider each class in turn.

Lemma 3.2. For $s \equiv_{3} 0$ and $s \equiv_{3} 1$ and $x>s+1$ :

$$
x \cdot 2= \begin{cases}x+1, & x-s \equiv_{2} 0 ; \\ x-1, & x-s \equiv_{2} 1 .\end{cases}
$$

Lemma 3.3. For $s \equiv_{3} 2$, and $x>s$,

$$
x \cdot 2= \begin{cases}x+2, & x-s \equiv_{4} 1,2 ; \\ x-2, & x-s \equiv_{4} 3,0 .\end{cases}
$$

## 4. Multiplication by 3

Multiplication by 3 is the last detailed case that is analyzed in this paper. Its structure is slightly more difficult than in the earlier cases. For each of the following lemmas, suppose that the seed is sufficiently large. The first lemma collects some preliminary calculations.

Lemma 4.1. $0 \cdot 3=2,1 \cdot 3=3,2 \cdot 3=4,3 \cdot 3=0,4 \cdot 3=1$.

Lemma 4.2. For $5 \leqslant x \leqslant s$ :

$$
x \cdot 3= \begin{cases}x+1, & x \equiv_{9} 5,8 \\ x+2, & x \equiv_{9} 6,1,2 \\ x-2, & x \equiv_{9} 7,0,4 \\ x-1, & x \equiv_{9} 3\end{cases}
$$

After each ninth step, the column becomes complete.
In the remainder of this section, only the 3 -column of the multiplication table for $s \equiv 32$ is considered, since this is the only case needed for the subsequent results. Note that $s \equiv 92,5,8$. Each case yields a different pattern after the row labelled by the seed.

Lemma 4.3. For $x>s \equiv_{9} 2$ :

$$
x \cdot 3= \begin{cases}x-2, & x-s \equiv_{4} 1,2 \\ x+2, & x-s \equiv_{4} 3,0\end{cases}
$$

Lemma 4.4. For $x>s \equiv_{9} 5$ :

$$
x \cdot 3= \begin{cases}x-1, & x-s \equiv_{2} 1 \\ x+1, & x-s \equiv_{2} 0\end{cases}
$$

Lemma 4.5. For $s \equiv{ }_{9} 8$, $(s+1) \cdot 3=s-1$. For $x \geqslant s+2$ :

$$
x \cdot 3= \begin{cases}x+1, & x-s \equiv_{2} 0 \\ x-1, & x-s \equiv_{2} 1\end{cases}
$$

## 5. Multiplication groups

In this section, the multiplication groups for each $Q_{s}$ are analyzed. The analysis yields easy proofs of some later theorems.

Consider

$$
G=\langle R(0), R(1), R(2)\rangle<\operatorname{Mlt}\left(Q_{s}\right)
$$

$$
\begin{aligned}
R(0) & =(0, s, s-1, s-2, \ldots, 1) \\
R(1) & =(s+1, s+2)(s+3, s+4) \ldots(s+2 n+1, s+2 n+2) \ldots \\
R(2) & =(0,2,1)(3,5,4) \ldots
\end{aligned}
$$

But one has to consider the seed mod 3 .

- For $s \equiv{ }_{3} 0$, one has $(0,2,1) \ldots(s-3, s-1, s-2) \cdot(s, s+1)(s+2, s+3) \ldots$
- For $s \equiv_{3} 1$, one has $(0,2,1) \ldots(s, s+1, s-1) \cdot(s+2, s+3) \ldots$
- For $s \equiv_{3} 2$, one has $(0,2,1) \ldots(s-1, s, s-2) \cdot(s+1, s+3)(s+2, s+$ 4) $(s+5, s+7)(s+6, s+8) \ldots$

Consider $R(0), R(1), R(2)$ in $S_{\mathbb{N}}$. A natural question is whether or not the groups

$$
G=\langle R(0), R(1), R(2)\rangle
$$

and

$$
F=\langle R(0), R(1), R(2), R(3)\rangle
$$

have transitive actions on $Q_{s}$. If so, are the groups multiply transitive?

### 5.1. Transitivity

Lemma 5.1. For all $s,\langle R(0)\rangle$ acts transitively on the hub.
Proof. By Lemma 2.2, $0 \cdot x=x-1$ for $0<x \leqslant s$ and $0 \cdot 0=s$. Thus $x R(0)^{x}=0$, and $0 R(0)^{y+1}=s-y$. Therefore for $x, z=s-y \in H$, there is an $n$ such that $x R(0)^{n}=z$.

Lemma 5.2. For $s \equiv{ }_{3} 0,1, Q_{s} \backslash H_{s}$ is in one orbit of the action of $G$ on $Q_{s}$. Moreover, one can choose $g \in G$ so that $x_{1} g=x_{2}$ for any $x_{1}, x_{2} \in Q_{s} \backslash H_{s}$ and $g$ stabilizes 1 .

Proof. Let $x=s+2 n-i, y=s+2 m-j$, where $n, m \in \mathbb{N}$ and $i, j \in\{0,1\}$. Let $\tau=R(1)^{i}(R(2) R(1))^{m-n} R(1)^{j}$. Now it is shown that $x \tau=y$. The initial multiplication by $R(1)^{i}$ sends both $s+2 n-i$ to $s+2 n$. Now by Lemmas 2.6 and 3.1 an application of $R(2) R(1)$ sends $s+2 n$ to $s+2 n+2$. So $(R(2) R(1))^{t}$ sends $s+2 n$ to $s+2 n+2 t$. Therefore $R(1)^{i}(R(2) R(1))^{t}$ sends $s+2 n-i$ to $s+2 n+2 t$. Finally $R(1)^{j}$ sends this to to $s+2 n+2 t-j$. Therefore $(s+2 n-i) \tau=s+2 n+2(m-n)-j=s+2 m-j$. To stabilize 1, use $\sigma=R(1)^{i} R_{1}(2,0)^{m-n} R(1)^{j}$. Note that since $R_{1}(2,0)=R(2) R(0) R(1)^{-1}$, on $Q_{s} \backslash H_{s}, R_{1}(2,0)$ behaves like $R(2) R(1)$, since $x R(0)=x$ and $x R(1)^{2}=x$ for $x \in Q_{s} \backslash H_{s}$. Thus $x \sigma=x R(1)^{i}(R(2) R(1))^{n-m} R(1)^{j}=y$ as above.

Theorem 5.3. The group $G$ acts transitively on $Q_{s}$ for $s \equiv_{3} 0,1$.

Proof. Using Lemmas 5.1 and 5.2, it remains to show that a hub element can be sent to a non-hub element. Note that $s \cdot 2=s+1$ in this case. So to send a hub element $h$ to a non-hub element $s+2 n-j$, use $\sigma=$ $R(0)^{h+1} R(2) R(1)(R(2) R(1))^{n-1} R(1)^{j}$.

For $s \equiv{ }_{3} 2$ the situation is more complex.
Lemma 5.4. Let $\sigma_{k, i}=R(2)^{k} R(1)^{i}$ for $k, i \in\{0,1\}$. Then in $Q_{s}$ for $s \equiv{ }_{3} 2$, $\sigma_{k, i}$ sends $s+4 n-2 k-i$ to $s+4 n$.

Proof. Since multiplication by 2 adds or subtracts $2, R(2)^{k}$ sends $s+4 n-$ $2 k-i$ to $s+4 n-i$. Now multiplication by 1 adds or subtracts 1 . So $R(1)^{i}$ sends $s+4 n-i$ to $s+4 n$.

Lemma 5.5. For $s \equiv_{9} 5,8, \tau=R(3) R(2) R(1)$ sends $s+4 n$ to $s+4 n+4$.
Proof. First, $(s+4 n) R(3)=s+4 n+1$ by Lemmas 4.4 and 4.5. Then $(s+4 n+1) R(2)=s+4 n+3$ by Lemma 3.3 and $(s+4 n+3) R(1)=s+4 n+4$ by Lemma 2.6. Thus ( $4 n) \tau=(4 n) R(3) R(2) R(1)=4 n+4$.

Lemma 5.6. For $s \equiv_{9}, 2 \tau=R(3) R(2)$ sends sends $s+4 n$ to $s+4 n+4$
Proof. First $(s+4 n) R(3)=(s+4 n+2)$ by Lemma 4.3. Then $(s+4 n)$ $(R(3) R(2))=s+4 n+4$ by Lemma 3.3.

Lemma 5.7. For $s \equiv_{3} 2, Q_{s} \backslash H_{s}$ is in one orbit of the action of $G$ on $Q_{s}$. Moreover, one can choose $g \in G$ so that $x_{1} g=x_{2}$ for any $x_{1}, x_{2} \in Q_{s} \backslash H_{s}$ and $g$ stabilizes 1 .

Proof. We show that any $x \in Q_{s} \backslash H_{s}$ can be sent to $y \in Q_{s} \backslash H_{s}$. Let $x=4 n-2 k-i$ and $y=4 m-2 k^{\prime}-i^{\prime}$, where $k, k^{\prime}, i,{ }^{\prime} i \in\{0,1\}$. Then for $\varphi=\sigma_{k, i} \tau^{m-n} \sigma k^{\prime}, i^{\prime-1}, x \varphi=y:$

$$
\begin{aligned}
(s+4 n-2 k-i) \varphi & =(s+4 n-2 k-i) \sigma_{k, i} \tau^{m-n} \sigma_{k^{\prime}, i^{\prime}} \\
& =(s+4 n) \tau^{m-n} \sigma_{k^{\prime}, i^{\prime}} \\
& =(s+4 m) \sigma_{k^{\prime}, i^{\prime}} \\
& =s+4 m-k^{\prime}-i^{\prime}
\end{aligned}
$$

Thus $x \varphi=y$. Note that outside the hub $R(0)$ stabilizes $x$. So $\alpha=$ $R_{1}(3,0) R_{1}(2,0) R(1)$ behaves like $R(3)$ and stabilizes 1 while $\beta=R_{1}(2,0) R(1)$ behaves like $R(2)$ and stabilizes 1 . Now apply Lemma 5.7 with $\alpha$ in place of $R(3)$ and $\beta$ in place of $R(2)$

Theorem 5.8. For $s \equiv_{3} 2, F$ acts transitively on $Q_{s}$.
Proof. It remains to be shown that one can send a hub element to a nonhub element as before. Let $h \in H_{s}$ and $x=s+4 n-2 k-i$. First, let $\psi=R(0)^{h+1} R(3) \sigma_{1,1} \tau^{n-1} \sigma_{k, i}$. Then $h \psi=x$ by the above lemmas.

### 5.2. 2-transitivity

The goal of this section is to prove that $\operatorname{Mlt}\left(Q_{s}\right)$ is 2-transitive.
Lemma 5.9. Let $H=\langle R(0), R(2)\rangle$. Then $H_{s}$ is in one orbital of the action of $H$ on $Q_{s}$ for $s \equiv{ }_{3} 0,1$.

Proof. Given $h_{1}, h_{2}, x_{1}, x_{2} \in H_{s}$, there is an $n$ so that $h_{1} R(0)^{n}=s$ (by Lemma 5.1). So $h_{1} R(0)^{n} R(2)=s+1$. Let $h_{2} R(0)^{n} R(2)=k$. Now choose $m$ so that $k R(0)^{m}=x_{2} R(0)^{-\left(s-x_{1}\right)} R(2)^{-1}$. Thus $h_{1} \sigma=x_{1}$ and $h_{2} \sigma=x_{2}$ for $\sigma=R(0)^{n} R(2) R(0)^{m} R(2)^{-1} R(0)^{s-x_{1}}$.

Lemma 5.10. Let $H=\langle R(0), R(3)\rangle$. Then $H_{s}$ is in one orbital of the action of $H$ on $Q_{s}$ for $s \equiv_{3} 2$.
Proof. Given $h_{1}, h_{2}, x_{1}, x_{2} \in H_{s}$, there is an $n$ so that $h_{1} R(0)^{n}=s$ (by Lemma 5.1). So $h_{1} R(0)^{n} R(3)=s+1$. Let $h_{2} R(0)^{n} R(3)=k$. Now choose $m$ so that $k R(0)^{m}=x_{2} R(0)^{-\left(s-x_{1}\right)} R(3)$. Thus $h_{1} \sigma=x_{1}$ and $h_{2} \sigma=x_{2}$ for $\sigma=R(0)^{n} R(3) R(0)^{m} R(3)^{-1} R(0)^{s-x_{1}}$.

Remark 5.11. The above two lemmas, along with the fact that $h R(1)=$ $h \forall h \in H_{s}$ show that the hub is in one orbital of the action of $F$.

Lemma 5.12. For $x_{1} \in Q_{s} \backslash H_{s}$ and $h_{1}, h_{2}, h_{3}$ there is a $\sigma$ so that $x_{1} \sigma=h_{2}$ and $h_{1} \sigma=h_{3}$.

Proof. Use $R(0)^{n}$ for some $n$ so send $h_{1}$ to 1. By Lemmas 5.2 and 5.7, there is a $\beta$ so that $1 \beta=1$ and $x_{1} \beta=s+1$. Then for $s \equiv_{3} 0,1 \gamma=R(0)^{n} \beta R(2)^{-1}$ is such that $x_{1} \gamma, h_{1} \gamma \in H_{s}$. For $s \equiv_{3} 2$ use $\gamma=R(0)^{n} \beta R(3)^{-1}$. Now since $H_{s}$ is in one orbital of the action of $\langle R(0), R(2), R(3)\rangle$ (Remark 5.11), the proof is complete.

Lemma 5.13. For $x_{1}, x_{2} \in Q_{s} \backslash H_{s}$ and $h_{1}, h_{2} \in H_{s}$, there is a $\sigma$ so that $x_{i} \sigma=h_{i}$.

Proof. Let $\alpha$ be so that $x_{1} \alpha=1$. Then perhaps $x_{2} \alpha=h \in H_{s}$. Then by Lemma 5.9 , there is a $\beta$, so that $1 \beta=h_{1}, h \beta=h_{2}$. Thus $\sigma=\alpha \beta$. If $x_{2} \alpha=x \notin H_{s}$ apply Lemma 5.12.

Theorem 5.14. F acts 2-transitively on $Q_{s}$.
Proof. We find a $\sigma$ that sends $\left(x_{1}, x_{2}\right) \in Q_{s}^{2}$ to $\left(y_{1}, y_{2}\right)$. First by the above three lemmas, there is a map $\alpha$ so that $\left(x_{1}, x_{2}\right) \alpha=(0,1)$, and a map $\beta$ so that $\left(y_{1}, y_{2}\right) \beta=(0,1)$. Then $\left(x_{1}, x_{2}\right) \alpha \beta^{-1}=\left(y_{1}, y_{2}\right)$

### 5.3. High transitivity

It has been shown how to construct permutations in $F \leqslant \operatorname{Mlt}\left(Q_{s}\right)$ that are 2-transitive. The question is whether one can go farther.

First note that since $F$ is 2-transitive it is primitive (Lemma 4.10 in [3]). Therefore we can apply Lemma 10.8 in [3] with the hub as the Jordan set. This theorem says that if a permutation group on $\Omega$ is primitive on an infinite set with a subgroup $H$ that is transitive on a set, $X$, and fixes the complement of $X$, the multiplication group is highly transitive. Moreover, if $X$ is finite, $\operatorname{Alt}(\Omega) \leqslant F$. Thus $\operatorname{Alt}(\mathbb{N}) \leqslant F \leqslant \operatorname{Mlt}\left(Q_{s}\right)$.

## 6. Subquasigroups

As noted in Remark 2.5, each greedy quasigroup has a unique singleton subquasigroup: $\{0\}$ in the elementary 2-group $Q_{0}$, and $\{1\}$ in $Q_{s}$ for $s>0$. We refer to the singleton subquasigroup and the empty subquasigroup as the trivial subquasigroups of the greedy quasigroups. The group $Q_{0}$ has uncountably many subquasigroups, since for each of the uncountably many subsets $S$ of $\mathbb{N}$, the vector

$$
\begin{equation*}
\left(0 \chi_{S}, 1 \chi_{S}, \ldots, n \chi_{S}, \ldots\right) \tag{4}
\end{equation*}
$$

of values of the characteristic function of $S$ generates a distinct subgroup of the isomorphic copy $(\mathbb{Z} / 2 \mathbb{Z})^{\mathbb{N}}$ of $Q_{0}$.

Proposition 6.1. The greedy quasigroup $Q_{1}$ has uncountably many subquasigroups.

Proof. Outside the hub $\{0,1\}$, the multiplication on $Q_{1}$ is constructed exactly as in $Q_{0}$. Thus for each subgroup $P$ of $Q_{0}$ with $\{0,1\} \leqslant P$, the subset $P$ of $\mathbb{N}$ forms a subquasigroup of $Q_{1}$. But $Q_{0}$ has uncountably many such subgroups $P$.

The respective hubs $H_{1}$ and $H_{2}$ of $Q_{1}$ and $Q_{2}$ form cyclic groups, with 1 as the identity element (Remark 2.5). These cases are exceptional.

Proposition 6.2. For $s>2$, there are no non-trivial subquasigroups of $Q_{s}$.
Proof. Note that $F$ is transitive for all $s \geqslant 3$. Thus if a subquasigroup, $H$, contains $0,1,2,3$ then $H=Q_{s}$. Let $H$ be a subquasigroup. If $0 \in H$, then $H_{s} \subset H$. In particular for $s \geqslant 3,0,1,2,3 \in H$ and $H=Q_{s}$. Suppose $x \neq 0,1 \in H$, then $x \cdot x=0 \in H$, so as above $H=Q_{s}$. Thus the only subquasigroup is the trivial subquasigroup $\{1\}$.

Proposition 6.3. For $s \geqslant 2, Q_{s}$ is simple.
Proof. This follows immediately since $\operatorname{Mlt}\left(Q_{s}\right)$ is 2-transitive.

## 7. Homomorphisms

Theorem 7.1. For $i \neq j, Q_{i} \not \not Q_{j}$.
Proof. In both $Q_{i}, Q_{j}, 0$ is the unique element that fixes infinitely many elements. So for any isomorphism $\varphi, \varphi: 0 \mapsto 0$. In $\operatorname{Mlt}\left(Q_{i}\right), R(0)$ is an $i+1$-cycle, but in $Q_{j} R(0)$ is a $j+1$-cycle. Thus $Q_{i} \not \neq Q_{j}$.

One can actually prove stronger results.
Lemma 7.2. Let $\varphi: Q_{i} \rightarrow Q_{j}$.
(a) If $\varphi$ is injective then there is a $k \in Q_{i}$ such that $k, k \varphi$ are both nilpotent.
(b) If $\varphi$ is surjective then there is a $k \in Q_{i}$ such that $k, k \varphi$ are both nilpotent.

Proof. There are only two elements $k \in Q_{i}$ such that $k \cdot k \neq 0$, namely 0,1 , and similarly for $Q_{j}$.
(a) Let $\varphi$ be injective. Suppose that $x \varphi, y \varphi$ are not nilpotent. Let $z$ be nilpotent, then $z \varphi$ is not $x \varphi, y \varphi$ and these are the only non-nilpotent elements in $Q_{j}$. Thus both $z, z \varphi$ are nilpotent.
(b) Since $\varphi$ is surjective, at most two of the nilpotent elements of $Q_{j}$ can be the image of non-nilpotent elements of $Q_{i}$. There must be nilpotent elements on $Q_{i}$ that are mapped to nilpotent elements of $Q_{j}$.

In what follows, the notations $q_{i}, q_{j}$ are used for an element $q \in Q_{i}$ to distinguish it from $q \in Q_{j}$.

Lemma 7.3. Let $\varphi: Q_{i} \rightarrow Q_{j}$ be a homomorphism and $0_{i} \varphi=0_{j}$. If $x \cdot x=0$, then $x \varphi \cdot x \varphi=0$.

Proof. $0_{j}=0_{i} \varphi=(x \cdot x) \varphi=x \varphi \cdot x \varphi$.
Lemma 7.4. Let $\varphi: Q_{i} \rightarrow Q_{j}$ be a homomorphism. If there is an element $x \in Q_{i}$ such that $x \cdot x=0$ and $x \varphi \cdot x \varphi=0$, then $0_{i} \varphi=0_{j}$.

Proof. Let $k$ be one such element. Then $0_{j}=0_{i} \varphi=(k \cdot k) \varphi=k \varphi \cdot k \varphi$.
Remark 7.5. In particular, Lemma 7.3 and Lemma 7.4 are true for surjective and injective homomorphisms.

Lemma 7.6. For any homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$ and $i, j \neq 0,1,1_{i} \varphi=$ $1_{j}$.

Proof. This follows from the fact that $1_{i}$ is the only idempotent element of $Q_{i}$. (Everything else other than $0_{i}$ is nilpotent).

Lemma 7.7. For any surjective (injective) homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$, $s_{i} \varphi=s_{j}$.

Proof. $s_{i} \varphi=\left(0_{i} \cdot{ }_{i} 0_{i}\right) \varphi=0_{i} \varphi \cdot{ }_{j} 0_{i} \varphi=0_{j} \cdot{ }_{j} 0_{j}=s_{j}$.
Remark 7.8. In fact, this is true if $0_{i} \varphi=0_{j}$.
Theorem 7.9 (Homomorphism Theorem). Suppose $i, j>1$.
(a) There is no injective homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$.
(b) There is no surjective homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$.

Proof. Note that by looking at the multiplication table for $Q_{j}$, that $s_{j} L\left(0_{j}\right)^{s_{j}}$ $=s_{j}$ and $s_{j} L\left(0_{i}\right)^{i} \neq s_{j}$ for $i<s_{j}$. Since $s_{i} \varphi=s_{j}$, then $s_{j}=s_{i} \varphi=$ $s_{i} R\left(0_{i}\right)^{i}=s_{i} \varphi R\left(0_{i} \varphi\right)^{i}=s_{j} R\left(0_{j}\right)^{i}$. Thus $j+1 \mid i+1$. Perhaps one can "loop" several times, but the loop must be completed. Thus there is no injective or surjective homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$, if $i<j$. So, suppose that $j+1 \mid i+1$, but $j \neq i$. Note that $s_{i} R(0)^{j-1}$ is nilpotent. Then $s_{i} R(0)^{j-1} \varphi=s_{i} \varphi R(0 \varphi)^{j-1}=s_{j} R\left(0_{j}\right)^{j-1}=1_{j}$. This is contradicts Lemma 7.3 , since a nilpotent must be mapped to a nilpotent and $1_{j}$ is idempotent.

Remark 7.10. Theorem 7.1 can be seen as a corollary to the Homomorphism Theorem.

Not only are the $Q_{i}$ 's are not isomorphic, there is no injective or surjective homomorphism between them. It is natural to ask whether there is any non-trivial homomorphism between them. Of course, there is the trivial homomorphism $x \varphi=1, \forall x \in Q_{i}$ for any $Q_{i}, Q_{j}$. It turns out that this is the only homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$ for $i \neq j$.

Theorem 7.11. The only homomorphism $\varphi: Q_{i} \rightarrow Q_{j}$ for $i \neq j$ is the trivial homomorphism.

Proof. Let $\varphi: Q_{i} \rightarrow Q_{j}$. If there is a nilpotent element $x$ such that $x \varphi$ is also nilpotent, by Lemma $7.40_{i} \varphi=0_{j}$, so then by Lemma $7.7 s_{i} \varphi=s_{j}$. Then the homomorphism fails as in Theorem 7.9. Thus for any nilpotent $x, x \varphi$ is either 0 or 1. If $x \neq 0$ and $x \varphi=0$, then $0 \varphi=(x \cdot x) \varphi=x \varphi x \varphi=0_{j} \cdot 0_{j}=s_{j}$. Then for any nilpotent $y, s_{j}=0 \varphi=(y \cdot y) \varphi=y \varphi \cdot y \varphi$. So $s_{j}$ is the square of $y \varphi$. Thus $y \varphi=0_{j}$ for any nilpotent $y$. Now, $s_{i} \varphi=\left(0_{i} \cdot 0_{i}\right) \varphi=0_{i} \varphi 0_{i} \varphi=$ $s_{j} \cdot s_{j}=0_{j}$. However, in any $Q_{i}$ there are nilpotent elements $x, y$ such that $x y=s_{i}$. Then $s_{i} \varphi=(x y) \varphi=x \varphi y \varphi=0_{j} \cdot 0_{j}=s_{j}$. This is a contradiction, so there is no $x$ so that $x \varphi=0_{j}$. Thus $x \varphi=1_{j}$ for all nilpotent $x$. In particular $s_{i} \varphi=1$, so $0_{i} \varphi=\left(s_{i} \cdot s_{i}\right) \varphi=s_{i} \varphi \cdot s_{i} \varphi=1_{j} \cdot 1_{j}=1$. Thus $\varphi$ is trivial.

Theorem 7.12. For $s>2$, there are only two endomorphisms of $Q_{s}$, the constant and the identity. In particular, $Q_{s}$ is rigid.

Proof. Suppose that $f: Q_{s} \rightarrow Q_{s}$ is an endomorphism. Since $Q_{s}$ is simple by Theorem 6.3, the kernel congruence of $f$ is either trivial (the equality relation) or improper. If it is improper, then $f$ is constant, its image being the unique singleton subquasigroup $\{1\}$ of $Q_{s}$. Otherwise, $f$ injects. Now 0 is the only element that is the square of more than one element, so $0 f=0$. The image $s f=(0 \cdot 0) f=0^{f} \cdot 0^{f}$ of the seed is a square, namely 1,0 or $s$. If $s f=1$, then $0^{f} \cdot 0^{f}=1$, yielding the contradiction $0 f=1$. Again, $s f=0$ would contradict the injectivity of $f$. Thus $s f=s$. By Lemma 5.2, $s-r=s R(0)^{r}$ for $0 \leqslant r<s$. Then $(s-r) f=s R(0)^{r} f=s R(0)^{r}=s-r$, so the hub is fixed. Since the hub generates all of $Q_{s}$, it follows that $Q_{s}$ is fixed, and $f$ is the identity.

## 8. Game theory applications

Greedy quasigroups are motivated in part by combinatorial games, in particular by nim. Nim is a game played with several piles, or heaps of counters. A player selects a pile and removes some, or possibly all the counters in the pile. The player to make the last move wins. With only two piles, the strategy is simple: equalize the piles, and then when your opponent removes $n$ counters from one pile, remove $n$ from the other. In this way, a player will never be at a loss for a move. With three or more non-empty piles, the strategy is a little more elusive. One must compute the nim-sum. The nim-sum is a way of reducing a collection of piles to a single value. This value represents the size of a single pile that is equivalent to the original position. If this pile were included in the original position, the resulting game would be a win for the first player. For details, see [1]. An alternative characterization of nim is that of a Rook on a quarter-infinite chessboard. Place a Rook on the board and make legal Rook moves up and left of the board. A player wins by placing the Rook on the upper-left corner. Now, greedy quasigroups have the following characterization as a game: place a nim-heap of size $n, n \geqslant 0$ on a chessboard. Move the heap as a Rook. Once the heap reaches the upper-left square, players may play in the nim heap. Clearly, with a single non-empty nim-heap on the board, one can win by forcing the other player to place the heap on the upper-left square and then removing the entire nim-heap. These game are examples of the sequential compounds in [9]. The difficulty arises when several heaps of different sizes are placed on the board. To play correctly, one must compute the value of each heap, with is a function of its size and its location. In this way, one can compute the nim-value of the position, and using combinatorial game theory, make the correct move. Suppose heaps of sizes $n_{1}, n_{2}, \ldots, n_{k}$ at locations $\left(x_{1}, y_{1}\right), \ldots,\left(x_{n}, y_{n}\right)$. The value of each heap is $x_{i} \cdot i y_{i}$ where $\cdot i$ is the multiplication in $Q_{n_{i}}$. The total value of the game is then:

$$
\bigoplus_{i=1}^{k} x_{i} \cdot y_{i}
$$

where $\oplus$ is nim-addition. A natural generalization is to place an entire game of nim on a square. This does not produce any new games, since each game of nim is equivalent to a single nim heap; so one might as well simply put the single nim-heap on the square.

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[^0]:    2000 Mathematics Subject Classification: 04A72
    Keywords: Lie algebra, interval-valued fuzzy set, Lie ideal, Noetherian Lie algebra.
    This research work was supported by PUCIT.

[^1]:    2000 Mathematics Subject Classification: 20N05
    Keywords: loops, inverse property
    This research was supported by National ICT Australia (NICTA). NICTA is funded through the Australian Government's Backing Australia's Ability initiative, in part through the Australian Research council.

[^2]:    ${ }^{1}$ http://users.rsise.anu.edu. au/ ~jks/IPloops/

[^3]:    2000 Mathematics Subject Classification: 05B15; 20N05
    Keywords: Quasigroup, loop, group, isotopy, translation

[^4]:    2000 Mathematics Subject Classification: 20N05, 08A05
    Keywords: central loops, central square, weak inverse property, cross inverse property, unique non-identity commutator, associator, square, Osborn loop.

[^5]:    2000 Mathematics Subject Classification: 20N15 05B15
    Keywords: irreducible quasigroups, latin hypercubes, n-ary quasigroups, reducibility, subquasigroup
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[^6]:    2000 Mathematics Subject Classification: 06F35, 03G25
    Keywords: S-absorbing set, (P)-decomposition, (P)-closed union, positive implicative hyper K-algebra, quasi union hyper K-algebra.

[^7]:    2000 Mathematics Subject Classification: 06F35, 03G25
    Keywords:(P)-closed union decomposition, union hyper K-algebra, positive implicative and implicative hyper K-algebra, S-absorbing, (P)-Decomposition

[^8]:    2000 Mathematics Subject Classification: 20N05, 91A46
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