ISSN 00002-3043

ЗШЗШОЅШЪР ФИШ SԵՂԵԿԱԳԻՐ ИЗВЕСТИЯ НАН АРМЕНИИ

UUBTUUSHUUMATEMATNKA

2017

Խ ՄԲԱԳՐԱԿԱՆ ԿՈԼԵԳԻԱ

Գլխավոր խմբագիր Ա. Ա. Մահակյան

Ն.Հ. Առաքելյան	Ռ. Վ. Համբարձումյան
Վ. Ս. Աթաբեկյան	Հ. Մ. Հայրապետյան
Գ.Գ. Գեորգյան	Ա. Հ. Հովհաննիսյան
Մ. Ս. Գինովյան	Վ. Ա. Մարտիրոսյան
Ն. Բ. Ենգիբարյան	Ք. Մ. Նահապետյան
Վ. Ս. Զաքարյան	Ք. Մ. Պողոսյան

Վ. Ս. Ջաքարյան

Ա. Ա. Թալալյան

Վ. Կ. Օհանյան (գլխավոր խմբագրի տեղակալ)

Պատասխանատու քարտուղար՝ Ն. Գ. Ահարոնյան

РЕДАКЦИОННАЯ КОЛЛЕГИЯ

Главный редактор А. А. Саакян

Г. М. Айрапетян	Н. Б. Енгибарян
Р. В. Амбарцумян	В. С. Закарян
Н. У. Аракелян	В. А. Мартиросян
В. С. Атабекян	Б. С. Нахапетян
Г. Г. Геворкян	А. О. Оганинсян
М С. Гиновян	Б. М. Погосян
В. К. Оганян (зам. главного редактора)	А. А. Талалян

Ответственный секретарь Н. Г. Агаронян

Известия НАН Арменни. Математика, том 52, п. 1. 2017, стр. 3-25.

THE PARTICLE STRUCTURE OF THE QUANTUM MECHANICAL BOSE AND FERMI GAS

A. BACH AND H. ZESSIN

Friedenstraße 12, D-48231 Warendorf, Germany,
Bielefeld University, Bielefeld, Germany.
E-mails: alexanderbach46@yahoo.de, zessin@math.un-bielefeld.de

Abstract. In the framework of von Neumann's description of measurements of discrete quantum observable we establish a one-to-one correspondence between symmetric statistical operators W of quantum mechanical systems and classical point processes κ_W, thereby giving a particle picture of indistinguishable quantum particles. This holds true under irreducibility assumptions if we fix the underlying complete orthonormal system. The method of the Campbell measure is developed for such statistical operators; it is shown that the Campbell measure of a statistical operator W coincides with the Campbell measure of the corresponding point process κ_W. Moreover, again under irreducibility assumptions, a symmetric statistical operator is completely determined by its Campbell measure. The method of the Campbell measure then is used to characterize Bose-Einstein and Fermi-Dirac statistical operators. This is an elementary introduction into the work of Fichtner and Freudenberg [10, 11] combined with the quantum mechanical investigations of [2] and the corresponding point process approach of [30]. It is based on the classical work of von Neumann [22], Segal, Cook and Chaiken [28, 8, 7] as well as Moyal [18].

MSC2010 numbers: 81-01; 82-01; 60G55

Keywords: Point process, Campbell measure, statistical operator.

1. INTRODUCTION

We consider quantum statistical states and ask for a precise particle picture of them. Under irreducibility assumptions we develop a one-to-one correspondence between symmetric statistical operators W of finite quantum mechanical systems and point processes κ_W , thereby giving a particle picture of indistinguishable quantum particles. This is done by developing a disintegration theory for such statistical operators in complete analogy to the decomposition of classical into conditional probabilities.

We also need the *method of the Campbell measure*, which is well known for point processes, and which is developed here for statistical operators. (This is inspired by the work of Fichtner, see for instance [12], and Liebscher [16].) We show that the Campbell measure of a symmetric statistical operator W coincides with the usual

A BACH AND H. ZESSIN

Campbell measure of its law κ_W , moreover, under irreducibility assumptions, W is then completely determined by its Campbell measure.

We then present the point processes which correspond to the quantum statistical operators of Maxwell-Boltzmann, Bose-Einstein and Fermi-Dirac in the case of a fixed number of particles. Surprisingly, only the point process belonging to the Maxwell-Boltzmann statistical operator is really known and has been considered in probability theory until now.

We then extend our considerations to systems with a random number of particles and therefore work on Fock spaces. In this framework the Poisson point process belongs to the Maxwell-Boltzmann statistical operator. Next the symmetric Bose-Einstein and Fermi-Dirac statistical operators are constructed together with their associated point processes. Since these statistical operators are determined by their Campbell measures, and since the Campbell measures coincide for statistical operators and their point processes, we shall investigate the Campbell measure of these point processes.

As a result of the application of the method of Campbell measures we find that the point processes belonging to Bose-Einstein and Fermi-Dirac statistical operators respectively are given by Papangelou processes with explicitly given conditional intensity kernels. They are called here *Polya sum* and *Pólya difference processes* respectively. The corresponding random fields are of first order and have independent increments. The distribution of the field variables, which represent the number of particles in a given region, are explicitly known. These results have been shown in [20]. Thus these processes have all characteristic properties of an ideal gas. In this way we obtain detailed informations about the point processes and thereby about the correponding statistical operators.

We stress here the point of view that for the developement of a full interacting theory of quantum gases one should start with the corresponding ideal gas and then modify this by means of a Boltzmann factor to include an interaction between the particles. (First steps in this direction can be found in [20].)

Historically the first attempts to unify quantum mechanics with point process theory can be found in the work of Fock [13], Segal [28], Cook [8] and Chaiken [7] and then, more systematically, in the work of Moyal [18]. For a more recent contribution to the construction of Bose and Fermi processes from the point of view of quantum mechanics we refer to Tamura and Ito [29].

Note added in February 2015. Unpublished versions of this work exist since 2008. We did not intend to publish it. But in the meantime several publications (see [20, 19, 26, 27] e.g.) referred to it so that it might be useful to make it available to the public.

2. DISINTEGRATION OF STATISTICAL OPERATORS

We consider von Neumann's description of the measuring process of discrete quantum observables (cf. [22, 23]) and use it for a representation of statistical operators in terms of their conditional statistical operators and their laws.

Consider a countable set $Y \neq \emptyset$ together with an equivalence relation \sim in Y. Represent (Y, \sim) by means of (Γ, r) in such a way that Γ is a countable set and $r: Y \to \Gamma$ a surjective mapping satisfying

$$(2.1) (x \sim y \Longleftrightarrow r(x) = r(y)).$$

Given $\gamma \in \Gamma$ we set $Y_{\gamma} = \{r = \gamma\}$ for the associated equivalence class. In the sequel we assume always that

$$(2.2) 1 \le \operatorname{cd} Y_{\gamma} < +\infty \quad \text{for any } \gamma.$$

Let $\mathcal H$ be a complex separable Hilbert space of countable dimension |Y|. We identify the set Y with the complete othonormal system (cons) $\mathcal Y=\{e_y|y\in Y\}$ chosen in $\mathcal H$. Furthermore, we set $\mathcal Y_+=\{e_y|y\in Y_\gamma\}$. The equivalence relation \sim induces an equivalence relation in $\mathcal Y$ by means of $(e_x\sim e_y\Leftrightarrow x\sim y)$ with $\mathcal Y_\gamma$ as equivalence classes.

The set of events of the system described by the Hilbert space $\mathcal H$ can be identified with the collection of all orthogonal projections resp. all (closed) subspaces. The state space $\mathcal S(\mathcal H)$ of the system is the collection of (self-adjoint) bounded linear operators $\mathcal W$ on $\mathcal H$ which are positive and have trace one, i.e. tr $\mathcal W=1$. Such $\mathcal W$ are called statistical operators. They form a convex set whose extremal points, the socialed pure states, are defined by

$$h \circ h = \langle h, . \rangle \cdot h, \qquad h \in \mathcal{H}, ||h|| = 1.$$

By the spectral theorem every state W admits a representation

$$\mathcal{W} = \sum_{n=1}^{\infty} p_n \cdot h_n \circ h_n,$$

where $(p_n)_n$ is a probability on N and $(h_n)_n$ some cons in \mathcal{H} . (For more details we refer to |9|.)

Our problem is how to associate to a given statistical operator $W \in \mathcal{S}(\mathcal{H})$, admitting a spectral resolution with respect to a given *cons* \mathcal{Y} , a law, and, in particular situations, a point process κ , and vice versa.

In the above situation we are given a complex separable Hilbert space $\mathcal H$ with fixed basis $\mathcal H$, indexed by Y. We consider

$$\mathcal{H}_{\gamma} = sp\{e_y|y \in Y_{\gamma}\},\$$

the smallest subspace of \mathcal{H} containing $\{e_y|y\in Y_\gamma\}$. The collection $(\mathcal{H}_\gamma)_{\gamma\in\Gamma}$ is an orthogonal decomposition of \mathcal{H} ; and \mathcal{H} is the direct sum of it. We have

$$1 \le \dim \mathcal{H}_{\gamma} = |Y_{\gamma}| = \operatorname{cd} Y_{\gamma} < \infty.$$

Here cd denotes cardinality. Finally we write

$$P_{\gamma} = P^{\mathfrak{H}_{\gamma}}$$

for the orthogonal projection onto \mathcal{H}_{γ} .

We start with a statistical operator $W \in \mathcal{S}(\mathcal{H})$ which admits the spectral resolution

(2.3)
$$\mathcal{W} = \sum_{y \in Y} P_y \, \varrho(y)$$

for some law ϱ on Y with respect to the chosen cons \mathcal{Y} . Here $P_y = e_y \diamond e_y$ with $e_y \diamond e_y = \langle e_y, . \rangle \cdot e_y$. Thus \mathcal{W} is diagonalized by the given cons \mathcal{Y} . Set

$$(2.4) W_{\gamma} = \sum_{y} P_{y} \varrho(y).$$

This defines self-adjoint linear operators on \mathcal{H}_{γ} leaving \mathcal{H}_{γ} invariant s.th.

$$W_{\gamma} = P_{\gamma} W P_{\gamma}, \qquad W_{\gamma} \mathcal{H}_{\gamma}^{\perp} = \{0\}.$$

Decomposition (2.4) is unique. If to $W_{\gamma} = tr(P_{\gamma}W)$ is strictly positive, we can normalize W_{γ} to obtain the following statistical operator on \mathcal{H} :

(2.5)
$$W(.|\gamma) = \frac{P_{\gamma}WP_{\gamma}}{\operatorname{tr}(P_{\gamma}W)}.$$

This is called the *conditional statistical operator of* W *given* P_{γ} . The notion of conditional statistical operators has been studied systematically by Cassinelli. Zanghi and Ozawa (cf. [6, 23] and the literature cited there).

Theorem 2.1. Given an equivalence relation in Y which can be represented by means of (Γ, r) in such a way that conditions (2.1) and (2.2) are satisfied, any statistical

operator $W \in S(\mathcal{H})$, admitting a spectral resolution (2.3) with respect to Y, can be represented as

(2.6)
$$W = \sum_{\gamma \in \Gamma} W(.|\gamma) \cdot \kappa_{W}(\gamma),$$

where $W(.|\gamma) \in S(\mathcal{H})$, leaving \mathcal{H}_{γ} invariant with $W(.|\gamma)\mathcal{H}_{\gamma}^{\perp} = \{0\}$, and where $\kappa_{\mathcal{W}}$ is a probability on Γ having the following properties:

(2.7)
$$\kappa_{\mathcal{W}}(\gamma) = tr(P_{\gamma}\mathcal{W}), \quad \gamma \in \Gamma.$$

This decomposition is unique.

In formula (2.6) and also later we use the convention that $W(1\gamma) \cdot v_W(\gamma) = 0$ if $\kappa_W(\gamma) = 0$. We call κ_W the law of the statistical operator W. It is some kind of partial trace of W with respect to γ , and we also write $\kappa_W(\gamma) = tr_{\gamma}(W)$. This means that $tr_{\gamma}(W) = \sum_{y \in Y_w} \langle e_y, We_y \rangle$. We observe that for the calculation of the law κ_W we can use the cons which is most convenient, because a trace does not depend on the choice of a cons. Decomposition (2.6) is completely analogous to the decomposition of classical probabilities into conditional probabilities; and it is the starting point for the solution of our problem.

3. DISINTEGRATION OF SYMMETRIC STATISTICAL OPERATORS

Consider next a finite group S acting on Y together with the equivalence relation \sim induced by S in Y by means of $x \sim y \iff \exists g \in S: y = g.x$. All orbits are finite, and S acts transitively on each of them. We assume also that (Y, \sim) is represented by (Γ, r) . As above $\mathcal H$ denotes a complex separable Hilbert space with a cons given by $\mathcal Y$. We consider then the unitary representation $\mathcal U = (\mathcal U_g)_{g \in S}$ induced by S on $\mathcal H$ by means of

$$\mathfrak{U}_0 h = \sum_y \lambda_y \cdot e_{gy}, \quad h = \sum_y \lambda_y e_y.$$

It is obvious that \mathcal{U} acts on \mathcal{H} as well as on each \mathcal{H}_{γ} . Thus each \mathcal{H}_{γ} as well as $\mathcal{H}_{\gamma}^{\perp}$ remains invariant under \mathcal{U} . The collection \mathcal{U}_{γ} of restrictions of \mathcal{U}_{g} , $g \in \mathcal{G}$, to the subspaces \mathcal{H}_{γ} is called an irreducible system, if any closed subspace S of \mathcal{H}_{γ} which remains invariant under \mathcal{U}_{γ} is either $\{0\}$ or \mathcal{H}_{γ} . This is equivalent to the condition that it does not commute with no non-trivial (self-adjoint) projection ([1], Exercise 1.3.D.) A statistical operator \mathcal{W} is called symmetric (with respect to \mathcal{G}) if

$$\mathfrak{U}_g \mathcal{W} \mathfrak{U}_{g^{-1}} = \mathcal{W} \text{ for any } g \in \mathfrak{G}.$$

In the sequel we consider symmetric W admitting a spectral resolution for $cons \ y$.

Lemma 3.1. W is symmetric if and only if each W, is symmetric.

Proof. By (3.1) combined with decomposition (2.6) W is symmetric iff

$$\sum_{\gamma} W_{\gamma} = \sum_{\gamma} U_{g} W_{\gamma} U_{g} , \text{ for any } g \in \mathcal{G}.$$

The uniqueness of the decomposition combined with the fact that each \mathcal{H}_{γ} resp. The remains invariant under \mathcal{U} immediately implies the result.

We need also the following result which in our context is Schur's lemma ([4], Satz 7.1 b.):

Lemma 3.2. Let W be symmetric. If the collection U_{γ} is irreducible then W_{γ} is of the form $W_{\gamma} = \kappa_{W}^{*}(\gamma) \cdot P_{\gamma}$. Here κ_{W}^{*} are non-negative functions on Γ , determined by the equation $\kappa_{W}^{*}(\gamma) = \langle e_{y}, We_{y} \rangle$, $y \in Y_{\gamma}$.

The positivity of k_W^* follows from the positivity of the statistical operator W. Thus we obtain the following disintegration of a symmetric statistical operator W.

Corollary 3.1. If W is symmetric and if each U, is irreducible then

$$\mathcal{W} = \sum_{\gamma \in \Gamma} \kappa_{\mathcal{W}}(\gamma) P_{\gamma} \quad \text{and} \quad \sum_{\gamma \in \Gamma} \kappa_{\mathcal{W}}(\gamma) \dim \mathcal{H}_{\gamma} = 1.$$

To summarize we have the following result.

Theorem 3.1. Under the assumption that each $U_{\gamma}, \gamma \in \Gamma$, is irreducible the equation

(3.2)
$$W = \sum_{\gamma \in \Gamma} \frac{1}{\dim \mathcal{H}_{\gamma}} P_{\gamma} \cdot \kappa(\gamma)$$

induces a one-to-one correspondence between symmetric statistical operators W on H, admitting a spectral resolution with respect to H, and probabilities κ on Γ .

This correspondence will be the main device in the sequel.

Corollary 3.2. If W is a symmetric statistical operator on \mathcal{H} , admitting a spectral resolution with respect to \mathcal{Y} , and if \mathcal{U}_{γ} is irreducible then the conditional statistical operator $\mathcal{W}(.|\gamma)$, if well defined, coincides with the normalized projection onto \mathcal{H}_{γ} :

(3.3)
$$W(.|\gamma) = \frac{1}{\dim \mathcal{H}_{\gamma}} \cdot P_{\gamma}.$$

Moreover, $\kappa_{W}(\gamma) = \dim \mathcal{H}_{\gamma} \cdot \kappa_{W}(\gamma), \gamma \in \Gamma$, the law of W, determines the operator W completely.

THE PARTICLE STRUCTURE OF THE QUANTUM

From now on the underlying group S is given by a finite symmetric group S(E) of all permutations σ of some finite set E. In this case we consider the following operators:

$$\Pi_{\pm} = \frac{1}{|E|!} \cdot \sum_{\sigma \in \mathcal{S}(E)} sgn_{\pm}(\sigma) \cdot \mathbb{U}_{\sigma}.$$

Here $sgn(\sigma) \in \{-1, +1\}$ denotes the sign of σ where sgn_+ is the identity and $sgn_- = sgn$. Both operators are orthogonal projections onto subspaces \mathcal{H}_+ and \mathcal{H}_- of \mathcal{H} and satisfy

(3.4)
$$\mathfrak{U}_{\sigma}\Pi_{+} = \Pi_{+}, \ \mathfrak{U}_{\sigma}\Pi_{-} = sgn(\sigma) \cdot \Pi_{-} \text{ for any } \sigma \in S(E).$$

In particular the operators Π_{+} and Π_{-} are symmetric. The elements of \mathcal{H}_{+} are also called symmetric; the elements of \mathcal{H}_{-} antisymmetric.

4. EXAMPLES

We consider the following standard finite setting (cf. [2, 24]). X is a finite, non-empty set of cardinality d: and $Y = X^n$. According to the convention of quantum mechanics the 1-particle space of a particle in X is given by \mathbb{C}^X , whereas the n-particle system is described by the complex Hilbert space $\mathcal{H} = \bigotimes^n \mathbb{C}^X$, i.e. the n-th tensor power of the 1-particle space. Note that \mathcal{H} coincides with \mathbb{C}^Y , and if n=0 then \mathcal{H} is the one-dimensional complex plane. In \mathbb{C}^X we choose some $cons\ (e_x)_{x\in X}$ conveniently. $\mathcal{H} = \{e_y = \mathcal{H}_{n-1}^{-1}e_x, y = (x_1, \dots, x_n) \in Y\}$ then is a cons in \mathcal{H} indexed by \mathcal{H} . If n=0 then \mathcal{H} is a singleton consisting of some unit vector $\mathbf{1}$ in \mathbb{C} fixed once and for all. The underlying symmetric group is given by the collection \mathcal{S}_n of bijections σ on $E = [n] = \{1, \dots, n\}$. \mathcal{S}_n acts on Y by means of

$$\sigma \longmapsto ((x_1, \dots, x_{\sigma^{-1}(1)}, \dots, x_{\sigma^{-1}(n)})).$$

It operates on H by means of the collection of unitary representations consisting of

$$U_{\sigma}: e_{\mathfrak{e}_1} \otimes \cdots \otimes e_{\mathfrak{e}_n} \longmapsto e_{\mathfrak{e}_{\sigma^{-1}(n)}} \otimes \cdots \otimes e_{\mathfrak{e}_{\sigma^{-1}(n)}}$$

and is then extended by linearity. We shall be interested in statistical operators which are symmetric, i.e. commute with the above representation of \mathcal{S}_n , and which admit a spectral resolution with respect to \mathcal{F} . Every observation \mathcal{W} of a system of identical particles has this property. The Hilbert spaces $\mathcal{H}_+, \mathcal{H}_-$, appropriate for the description of particles obeying quantum statistics, are constructed by means of the projections Π_+, Π_- induced by the group \mathcal{S}_n .

A representation (Γ, r) of the equivalence relation induced by S_n on Y is given by

$$\Gamma = M_n(X) := \{\delta_{x_1} + \dots + \delta_{x_n} | (x_1, \dots, x_n) \in Y\},$$

$$r : (x_1, \dots, x_n) \longmapsto \delta_{x_n} + \dots + \delta_{x_n}.$$

4.1. The Maxwell-Boltzmann statistical operator. In \mathcal{H} we choose a consindexed by Y in the following way: We are given a statistical operator w on the 1-particle space $\mathcal{H}_1 := \mathbb{C}^X$. Denote by ϱ the probability on X appearing in the spectral resolution of w, which at the same time gives a cons $(e_x)_{x \in X}$ in \mathcal{H}_1 . This basis will be fixed also in the following examples and enables one to define the cons Y in X as above. Moreover, we always assume that ϱ is not a Dirac measure. This implies that $d = \operatorname{cd} X \geq 2$. The Maxwell-Boltzmann statistical operator for w is defined by the tensor product of w: $M^n = w^n$. Here w^n denotes the n-fold tensor product of w. Using proposition 16.3. in [24] this statistical operator can be expressed explicitly by

$$M_{in}^{n} = \sum_{y \in Y} P_{ij} \cdot \varrho^{n}(y),$$

where $P_y = \epsilon_y \circ e_y$, and ϱ^n is the product law $\varrho \otimes \cdots \otimes \varrho$ on Y. (4.1) is nothing else than the spectral resolution of \mathbb{M}^n with respect to \mathcal{Y} . \mathbb{M}^n_w is symmetric with respect to \mathcal{S}_n . By Theorem 2.1 there is associated the following law on $\mathbb{M}_n(X)$, which thus is a point process in X, namely

(4.2)
$$\kappa(\gamma) = \binom{n}{\gamma} \cdot \prod_{x \in X} \varrho(x)^{\gamma(x)}, \qquad \gamma \in \mathcal{M}_n(X).$$

Here '

$$\binom{n}{\gamma} = \frac{n!}{\prod_{x \in X} \gamma(x)!}, \qquad \gamma \in \mathcal{M}_n(X).$$

(4.2) follows from the fact that dim $\mathcal{H}_{\gamma}^{n} = \binom{n}{\gamma}$ and that, for $y = (x_1, \dots, x_n) \in Y_{\gamma}$ and thereby $\gamma = \delta_{x_1} + \dots + \delta_{x_n}$, by formula (4.1),

$$\kappa^*(\gamma) = \langle e_{x_1} \otimes \cdots \otimes e_{x_n}, \mathbb{M}_w^n e_{x_1} \otimes \cdots \otimes e_{x_n} \rangle = \prod_{j=1}^n \varrho(x_j).$$

The point process κ is called *Maxwell-Boltzmann process* for the parameters (ϱ, n) , and will be denoted by \mathbb{P}^n .

4.2. The Bose-Einstein statistical operator. We start with the following observations: We are given a particle number $n \ge 0$. One can construct by means of \mathcal{Y} , as chosen

above, a cons y+ in M and y- in H- respectively as follows:

$$\begin{split} & \forall_{+} = \Big\{ e_{+}(\gamma) = \sqrt{\binom{n}{\gamma}} \cdot H_{+} \otimes_{a \in supp \gamma} e_{a}^{\otimes \gamma(a)} | \gamma \in \mathcal{M}_{n}(X) \Big\}, \\ & \forall_{-} = \Big\{ e_{-}(\gamma) = \sqrt{n!} \cdot H_{-} \otimes_{a \in supp \gamma} e_{a} | \gamma \in \mathcal{M}_{n}(X) \Big\}. \end{split}$$

Here the tensor product is taken along a fixed numeration of X, and

$$\mathcal{M}_n(X) = \left\{ \delta_{x_1} + \cdots + \delta_{x_n} | (x_1, \dots, x_n) \in Y \right\}.$$

Y is the collection of all $y \in Y$ with pairwise distinct components

We work separately in each of the spaces M_{\pm} with these cons. In terms of \mathcal{Y}_{\pm} the projections H_{\pm} can be written as $H_{\pm} = \sum_{e \in \mathbb{N}_{\pm}} Q_e^{\pm}$, where the one-dimensional projections are given by $Q_e^{\pm} = e \diamond e$, $e \in \mathcal{Y}_{\pm}$. Since there is a bijection between $M_{\pi}(X)$ and \mathcal{Y}_{\pm} resp. $M_{\pi}(X)$ and \mathcal{Y}_{\pm} we see immediately that (recall that d = trX)

$$\operatorname{cd} \mathcal{Y}_+ = \binom{d+n-1}{n}; \quad \operatorname{cd} \mathcal{Y}_- = \binom{d}{n} \quad \text{if} \quad n \leq d; \quad \operatorname{cd} \mathcal{Y}_- = 0. \quad \text{if} \quad n > d.$$

The Bose-Einstein statistical operator for w is given by the conditional Maxwell-Boltzmann statistical operator given the projection H_+ . This is an operator on \mathcal{H}_+ defined by $\mathbb{E}^n = \frac{1}{tr(H_+M_-^n)} \cdot H_+M_-^n$. Note that

$$tr(H_+\mathbb{M}^n_w) = \sum_{u \in X} \prod_{a \in X} \varrho(a)^{\mu(u)} > 0.$$

because the ϱ is assumed not to be a Dirac measure.

We choose a cons in \mathcal{H}_+ for which \mathbb{E}^n can be diagonalized, namely \mathcal{Y}_+ , which is indexed by the finite set $\mathcal{M}_n(X)$. The symmetric group now acts on the basis \mathcal{Y}_+ and is trivial, i.e. a singleton consisting of the identity. Thus the associated equivalence relation \sim is given by the identity of elements in \mathcal{Y}_+ ; and the representation of (\mathcal{Y}_+, \sim) is given by $\Gamma = \mathcal{M}_n(X)$ with $r: e_+(\gamma) \longrightarrow \gamma$. Theorem 3.1 then implies that the point process belonging to \mathbb{E}^n is given by the following point process in X: For any $\gamma \in \mathcal{M}_n^n(X)$

(4.3)
$$\mathsf{E}_{n}(\gamma) = \frac{1}{\sum_{\mu \in \mathcal{M}_{n}(X)} \prod_{a \in X} \varrho(a)^{\mu(a)}} \cdot \prod_{a \in X} \varrho(a)^{\gamma(a)}.$$

Morcover, the Bose-Einstein statistical operator admits the representation

$$\mathbb{E}_w^n = \sum_{\gamma \in \mathcal{M}_n(X)} \mathsf{E}_v^n(\gamma) \cdot Q_{e_+(\gamma)}^+$$

We call E^n the Bose-Einstein point process in X for the parameters (n,ϱ)

If ϱ is the uniform distribution on X, and thereby $w=\frac{1}{d}trI$, where I denotes the identity operator on \mathbb{C}^N , then

$$\mathbb{E}_w^n := \mathbb{E}_d^n := \frac{1}{\left(\begin{array}{c} d+n-1 \\ n \end{array}\right)} \cdot \Pi_+,$$

and the Bose-Einstein process is then given by the uniform distribution on $\mathcal{M}_{\mathbb{Z}}(X)$:

$$\mathbb{E}_d^n(\gamma) = \frac{1}{\binom{d+n-1}{n}}$$
, $\gamma \in \mathcal{M}_n(X)$.

4.3. The Fermi-Dirac statistical operator. For $n \leq d = \operatorname{cd} X$ the Fermi-Dirac statistical operator for w is given by the conditional Maxwell-Boltzmann statistical operator given the projection H_{-} . This is a symmetric statistical operator on \mathcal{H}_{-} defined by

$$\mathbb{D}_w^n = \frac{1}{tr(\Pi_- \mathbb{M}_w^n)} \cdot \Pi_- \mathbb{M}_w^n$$

This operator admits a spectral resolution with respect to the $cons \, \mathcal{Y}_-$ in \mathcal{H}_- , where again the basis $(e_x)_{x \in X}$ is coming from the spectral resolution of w and ϱ is the corresponding law not being a Dirac measure. By Theorem 3.1 we then obtain as before the particle picture of the Fermi-Dirac statistical operator: It is given by the following simple point process, called Fermi-Dirac process for (n, ϱ) in X:

(4.5)
$$D_n^n(\gamma) = \frac{1}{Z} \operatorname{tr} \prod_{a \in X} \varrho(a)^{\gamma(a)}, \qquad \gamma \in \mathcal{M}_n(X), \text{ and } 0 \text{ otherwise }.$$

The partition function now is given by $Z = \sum_{n \in \mathcal{N}_{+}(X)} \prod_{a \in X} n(a)^{n-1}$. Thus D_{ϱ}^{n} is the conditional law of E_{ϱ} given $\mathcal{M}_{n}(X)$, i.e. given that the realization γ of the particle process is simple. We again have a representation of the Fermi-Dirac statistical operator which is parallel to the one for the Bose-Einstein statistical operator, namely

$$\mathbb{D}^n_w = \sum_{\gamma \in \mathcal{M}_{\mathbb{Z}}(X)} \mathsf{D}^n_{\varrho}(\gamma) \cdot Q^-_{r_{-1}(\gamma)}$$

Note that in the special case where $w = \frac{1}{d} \cdot I$, thus ϱ being the uniform distribution on X, the Fermi-Dirac statistical operator is given by

$$\mathbb{D}_w^n := \mathbb{D}_d^n := \frac{1}{\left(\begin{array}{c} d \\ n \end{array}\right)} \cdot \Pi_-,$$

and the simple point process by the Fermi-Dirac process in X for the parameters (n,d). (Recall that d=|X|.)

$$D_d^n(\gamma) = \frac{1}{\binom{d}{n}}, \quad \gamma \in M_n(X)$$

5. THE METHOD OF THE CAMPBELL MEASURE

In the situation of the last section we introduce the occupation number operator and the Campbell operator respectively Campbell measure of a statistical state.

The situation is the same as in the examples: $\mathcal{H}_1 = \mathbb{C}^X$ for some finite X; $(e_x)_{x \in X}$ is a cons in \mathcal{H}_1 . Recall that $\Gamma = \mathcal{M}_n^+(X)$, and $r: (x_1, \ldots, x_n) \longmapsto \delta_{x_1} + \cdots + \delta_{x_n}$. Note that $r = M \circ \iota$, where $\iota: (x_1, \ldots, x_n) \longmapsto e_{x_1} \otimes \cdots \otimes e_{x_n}$ and $M(e_{x_1} \otimes \cdots \otimes e_{x_n}) = \delta_{x_1} + \cdots + \delta_{x_n}$.

We define for $x \in X$ the occupation number operator in x on $\mathcal{H} = \mathcal{H}_1^{\otimes n}$ as follows: If I is the identity operator on \mathcal{H}_1 , let

(5.1)
$$N_x = \sum_{j=1}^n I \otimes \cdots \otimes \underbrace{e_x \circ e_x}_{j} \otimes \cdots \otimes I.$$

(In the case n=0 we set $\mathcal{N}_x^{(0)}=0\cdot I$.) And, more generally, $\mathcal{N}_B=\sum_{x\in B}\mathcal{N}_x$ the occupation number operator in $B\subset X$. It is evident that $\mathcal{N}_B=\zeta_B(M)I^n$, where for $x_1,\ldots,x_n\in X$ we set

$$\zeta_B(\delta_{x_1}+\cdots+\delta_{x_n})=(\delta_{x_1}+\cdots+\delta_{x_n})(B).$$

Extend $\mathcal{N}_{(1)}$ linearly to an operator-valued measure on $X \times \mathcal{M}^{\circ}(X)$ by $\mathcal{N}_h = \zeta_h(M) \cdot I^n$, $h \in F_+(X \times \mathcal{M}^{\circ}(X))$. Here $\zeta_h(\mu) = \int h(x,\mu) \, \mu(\mathrm{d}\,x)$, and F_+ denotes the collection of non-negative, measurable functions on the underlying domain. Thus in particular $\mathcal{N}_{B \times C} = \zeta_B(M) \cdot I_C(M) \cdot I^{\otimes n}$. This shows: Any element $e_y = 0$ of the basis is an eigenvector of $\mathcal{N}_{B \times C}$ with eigenvalue $\zeta_B(M(e_y)) \cdot I_C(M(e_y))$.

We are now in the position to define the Campbell measure for statistical operators on \mathcal{H} . Given a statistical operator W we call WN on \mathcal{H} the Campbell operator measure of W. Its trace $\mathcal{C}_{W}(.) = tr(WN_{(.)})$ is called the Campbell measure of W on $X \times M$. (X). Recall that the Campbell measure of the law κ_{W} of W is defined by

$$\mathfrak{C}_{\kappa_{\mathcal{W}}}(a,\gamma) = \gamma(a)\kappa_{\mathcal{W}}(\gamma), \qquad a \in X, \gamma \in \mathcal{M}^+(X).$$

It is obvious that such a Campbell measure is supported by the set $\{(a, \gamma) : \gamma(a) \geq 1\}$. Moreover, we see that the law $\kappa_{\mathcal{W}}$ of \mathcal{W} is determined by its Campbell measure.

Proposition 5.1. For any statistical operator W on the space H its Campbell measure coincides with the Campbell measure of its law, i.e. $\mathfrak{C}_W = \mathfrak{C}_{\kappa_W}$. The law of W is completely determined by \mathfrak{C}_W . If W is also symmetric then, under the additional irreducibility assumptions of Theorem 2. even W is completely determined by its Campbell measure.

Proof.

$$tr(\mathcal{WN}_h) = \sum_{y} \langle e_y, \mathcal{WN}_h(e_y) \rangle = \sum_{y} \hat{\varsigma_h}(r(y)) \langle e_y, \mathcal{W}(e_y) \rangle = \sum_{\gamma} \hat{\varsigma_h}(\gamma) \sum_{y \in Y_{\gamma}} \langle e_y, \mathcal{W}(e_y) \rangle.$$

The assertion now follows from the definition (2.7) of $\kappa_{\mathcal{W}}$. The remaining statement follows immediately from Theorem 3.1.

We remark for later use that Proposition 5.1 remains true for statistical operators W acting on subspaces of \mathcal{H} because the occupation number operators N_H act on them by restriction

6. STATES ON FOCK SPACES AND THEIR CAMPBELL MEASURES

The above picture is now extended to systems with a random particle number.

Let X be a finite set of cardinality $d \geq 1$ and $\mathcal{X}_m = \bigotimes^m \mathbb{C}^X$, $m \geq 0$, with $\mathcal{H}_0 = \mathbb{C}$. The cons in \mathbb{C} consists of some unit vector, denoted by 1. The direct sum of these Hilbert spaces is the Fock space over \mathbb{C}^X , denoted by \mathbb{H} . For each m the symmetric group \mathcal{S}_m acts on X^m , and the corresponding unitary representation on \mathcal{H}_m is denoted by \mathcal{U}_m . This family of representations gives rise to a unitary operator \mathcal{U} on \mathbb{H} , defined by the direct sum $\mathcal{U} = \sum_{m=0}^\infty \mathcal{U}_m$. Thus $\mathcal{U}(g)h = \mathcal{U}_m(g)h$, if $g \in \mathcal{S}_m, h \in \mathcal{H}_m$. Given statistical operators \mathcal{W}_m on \mathcal{H}_m and scalars $p_m \geq 0$, $m \geq 0$, summing up to 1, then the direct sum

$$(6.1) \mathcal{W} = \sum_{m=0}^{\infty} p_m \mathcal{W}_m$$

is a statistical operator on the Fock space \mathbb{H} . W is symmetric if and only if each \mathcal{W}_m has this property. It is obvious that the point process belonging to this statistical operator is given by

(6.2)
$$\kappa_{\mathcal{W}} = \sum_{m=0}^{\infty} p_m \cdot \kappa_{\mathcal{W}_m}$$

The simplest examples are obtained if $\mathcal{W}_m = w^m$ for some given statistical operator w on $\mathcal{H}_1 = \mathbb{C}^X$. Only them will be considered in the sequel in detail. In this framework the occupation number operator is given by the direct sum operator $\mathcal{N}_x = \sum_{m=0}^\infty \mathcal{N}_r$ on the Fock space over \mathbb{C}^X . Here

as defined above. And again $\mathcal{N}_B = \zeta_B(M) \cdot I$, $B \subset X$, where I now denotes the identity operator on \mathbb{H} . Extending \mathbb{N} to an operator valued measure on $X \times \mathbb{M}$ (X) as above by $\mathcal{N}_h = \zeta_h(M) \cdot I$, $h \in F_+(X \times \mathbb{M} \setminus X)$, we are now in the position to define the Campbell measure for statistical operators on \mathbb{H} as we did already in a special situation. Recall that $\chi_h(\mu) = \int h(x,\mu) \, \mu(\mathrm{d}\,x)$.

Given a statistical operator W on \mathbb{H} we call $WN_{(.)}$ the Campbell operator measure of W. By Theorem 2.1 we know that $WN_h = \sum_{\gamma \in \Gamma} \zeta_h(\gamma) \kappa_W(\gamma) \cdot W(|\gamma)$, $h \in F_+$. Define $\mathcal{C}_W(.) = tr(WN_{(.)})$. This object is called the Campbell measure of W. Arguing as above we obtain

Theorem 6.1. For any statistical operator W on the Fock space \mathbb{H} one has $\mathcal{C}_{W} = \mathcal{C}_{\kappa_{W}}$. Thus the law of W is completely determined by \mathcal{C}_{W} . If W is also symmetric then, under the additional irreducibility assumptions of Theorem 3.1, even W is completely determined by its Campbell measure.

Consider now the direct sums $\Pi_{\pm} = \sum_{n=0}^{\infty} \Pi_{\pm}^{(n)}$, where $\Pi_{\pm}^{(n)}$ is the orthogonal projection onto the *BE*- resp. *FD symmetric subspace* of \mathcal{H}_{m} . Π_{\pm} is then the orthogonal projection onto the *BE*- resp. *FD symmetric subspace* \mathbb{H}_{\pm} of \mathbb{H} . It follows (see [2]) that Π_{\pm} satisfy

(6.3)
$$\mathbf{u}_{\sigma}\Pi_{\pm} = sgn_{\pm}(\sigma)\Pi_{\pm}, \quad \sigma \in \mathcal{S}_{\infty} := \bigcup_{m \geq 0} \mathcal{S}_{m}.$$

We are mainly interested in statistical operators W living on the symmetric subspaces \mathbb{H}_{\pm} . By this we mean that W satisfies the conditions $W = \Pi_{\pm}W\Pi_{\pm}$. In case + this is equivalent to say that W is Bose-Einstein symmetric, i.e. $\mathbb{U}_{\sigma}W = W$. $\sigma \in \mathcal{S}_{\infty}$: and in case – that W is Fermi-Dirac symmetric, i.e. $\mathbb{U}_{\sigma}W = sgn(\sigma)W$. $\sigma \in \mathcal{S}_{\infty}$. Moreover, these conditions imply the symmetry of the statistical operator. (All this can be found in [2])

Theorem 6.1 remains true for statistical operators acting on the Fock spaces \mathbb{H}_{\pm} because the N_B act on \mathbb{H}_{\pm} by restriction. Note also that one obtains by means of a basis in \mathcal{H}_1 a basis in the Fockspaces \mathbb{H} . \mathbb{H}_{\pm} by taking unions $\bigcup_{m\geq 1} \bigcup_{m\geq 1} \bigcup$

7. STATES WITH RANDOM PARTICLE NUMBERS

The method of second quantization is recalled which permits to lift an operator on a 1-particle space to a Fock space.

7.1. The method of second quantization. We recall the method of the so-called second quantization. The idea behind is to lift operators H on $\mathcal H$ to one of the Fock spaces. The method goes back to the work of Fock [13], Cook [8] and Berezin [3] (cf. also [5]). If H is a statistical operator on $\mathcal H$, one can define a operator H_m on the tensor product $\mathcal H_m$ by setting $H_0 \mathbf 1 = 0$ and

$$H_m(e_{a_1} \otimes \cdots \otimes e_{a_m}) = \sum_{j=1}^m e_{a_1} \otimes \cdots \otimes H(e_{a_j} \otimes \cdots \otimes e_{a_m}), \quad a_1, \ldots, a_m \in X$$

Denoting by δ_{ik} the Kronecker symbol,

$$H_m = \sum_{i=1}^m H^{\delta_{jn}} \otimes \cdots \otimes H^{\delta_{jm}}$$

The direct sum of the H_m is denoted by

$$d\Gamma(H) = \sum_{m=0}^{\infty} H_m.$$

Note that we used this method already for the operator $e_x \circ e_x$ and obtained in chapter 6 for the operator $d\Gamma(e_x \circ e_x)$ the occupation number operator N_x on the Fock space over \mathbb{C}^X .

If w is a statistical operator on \mathcal{H} , the second quantization of w then is defined by

$$\Gamma(w) = \sum_{m=0}^{\infty} \frac{1}{m!} w^m$$

This is an operator on the full Fock space H having finite trace e.

An important observation is given in terms of such trace class operators. These are multiples of statistical operators, i.e. operators of the form $w = z \overline{w}^m$, where z > 0 and w is some statistical operator. In this case

$$\Gamma(w) = \sum_{m=0}^{\infty} \frac{z^m}{m!} \cdot w^m \text{ with } tr \, \Gamma(w) = e^z$$
 .

Lemma 7.1. Let II be a bounded, self adjoint operator such that $w = \exp(-\beta II)$ is a trace class operator with $\beta \in \mathbb{R}_+$. Then

$$\exp (-\beta H)^m = \exp \left(-\beta \sum_{j=1}^m H^{\delta_{j1}} \otimes \cdots \otimes H^{\delta_{jm}}\right)$$

Recall here that the left hand side of this equation is given by $e^{-\beta H} \otimes \cdots \otimes e^{-\beta H}$. For a proof of the lemma we refer to Cook [8].

Lemma 7.2. Let II be a self adjoint operator such that $w = \exp(-\beta H)$ is a trace class operator with $\beta \in \mathbb{R}$. Defining the associated Gibbs state

(7.1)
$$G = \frac{1}{tr \exp(-\beta H)} \exp(-\beta H)$$

and $z = cd \exp(-\beta H)$ we obtain

(7.2)
$$\Gamma(\exp(-\beta H)) = \sum_{m=0}^{\infty} \frac{e^m}{m!} G^m$$

 $\Gamma\left(\exp(-\beta H)\right)$ is trace class with trace e^2 .

As a consequence we see that $\mathbb{M}_{zG} := e^{-z} \Gamma(\exp(-\beta H))$ is a statistical operator on the Fock space.

According to Lemmas 7.1 and 7.2 there are two representations of this operator:

$$M_{\sigma G} = e^{-z} \sum_{m=0}^{\infty} \frac{z^m}{m!} G^m = e^{-z} \sum_{m=0}^{\infty} \frac{1}{m!} \exp \left(-\beta \sum_{j=1}^m H^{\delta_{j1}} \otimes \cdots \otimes H^{\delta_{jm}}\right).$$

To summarize in a slightly modified way: Given some trace class operator w = zw with corresponding spectral measure $\varrho = zw$ then w^m has trace tr $w^m = z^m$. In this case the associated second quantization of w is given by

(7.3)
$$M_{w} = \frac{1}{\Xi_{w}} \sum_{m=0}^{\infty} \frac{tr \, w^{m}}{m!} \cdot \frac{w^{m}}{cd \, w^{m}} = \frac{1}{\Xi_{w}} \sum_{m=0}^{\infty} \frac{z^{m}}{m!} \cdot \widetilde{w^{m}}.$$

 Ξ_w is the normalizing constant. In this way the trace class operator w is lifted to some symmetric statistical operator on the full Fock space \mathbb{H} .

The construction principle behind the *method of second quantization* is: Given m, the trace class operator w^m is normalized to some statistical operator w^m , then weighted by the factor $\frac{cd w}{m!}$ and summed up: finally it is normalized so that the resulting operator becomes statistical.

One also uses this quantization method in a slightly generalized form to lift the underlying w on the subspaces \mathbb{H}_{+} and obtain the statistical operators

$$\mathbb{E}_{w} = \frac{1}{\Xi^{+}} \sum_{m=1}^{\infty} tr(\Pi_{+}^{(m)} w^{m}) \cdot \frac{\Pi_{+}^{(m)}}{tr(\Pi_{+}^{(m)} w^{m})}$$

$$\mathbb{D}_{w} = \frac{1}{\Xi_{w}} \sum_{m=0}^{\infty} tr(\Pi_{-}^{(m)} w^{m}) \cdot \frac{\Pi_{-}^{(m)} w^{m}}{tr(\Pi_{-}^{(m)} w^{m})}.$$

Note here that the normalizing constants $\Xi^{\pm} = \sum_{m=0}^{\infty} (H_{-}^{-}M_{-})$ are termwise strictly positive and convergent on account of the assumption that ϱ is not a Dirac measure. \mathbb{E}_{w} is called the Bose-Einstein operator for w, \mathbb{D}_{w} the Fermi-Dirac operator

for w and $p_{g}^{\pm}: m \longmapsto \frac{1}{\pm 1} \cdot tr(\Pi_{\pm}^{(m)} \mathbb{M}_{m}^{m})$ the particle number distribution of \mathbb{E}_{w} or \mathbb{E}_{w} respectively. Thus the operators \mathbb{M}_{w} , \mathbb{E}_{w} and \mathbb{D}_{w} are the second quantizations of w for the different Fock spaces \mathbb{H} , \mathbb{H}_{\pm} . One question then is to calculate the corresponding laws and to characterize them.

7.2. Maxwell-Boltzmann statistical operators with Poissonian random particle number. The Maxwell-Boltzmann statistical operator is described as a solution of an integration-by-parts formula.

We are in the framework of section 4: \overline{w} is a statistical operator on \mathbb{C}^X . X being a finite set of cardinality d. As above we choose a cons $e_x, x \in X$, the one coming from the spectral decomposition of w with law o. We are interested in the symmetric statistical operator given by the second quantization of the trace class operator $w = \frac{1}{2} w^2$.

(7.4)
$$M_w = e^{-z} \sum_{m=0}^{\infty} \frac{z^m}{m!} \cdot M_w^m$$

This is the Maxwell-Boltzmann statistical operator for z, w. We remark that, instead of the Poisson law, any law $(p_m)_m$ can be taken to get some statistical operator. By formula (6.2) the corresponding point process is the Poisson process P_θ with intensity $\rho = z\bar{\varrho}$. Thus $p_m = P_\theta$, where

$$P_{\boldsymbol{\theta}}(\varphi) = \kappa_{M_m}(\varphi) = \mathrm{e}^{-z} \sum_{m=0}^{\infty} \frac{z^m}{m!} \sum_{(x_1, \dots, x_m) \in X^m} \varphi(\delta_{x_1} + \dots + \delta_{x_m}) \ \bar{\varrho}(x_1) \dots \bar{\varrho}(x_m).$$

 P_{y} is supported by $\mathcal{M}^{\circ}(X) = \bigcup_{n=0}^{\infty} \mathcal{M}^{\circ}(X)$. Note that this formula is completely parallel to (7.4), namely

$$\kappa_{\mathcal{M}_w} = e^{-z} \sum_{m=0}^{\infty} \frac{z^m}{m!} (L_{\hat{\theta}})^{*i\hat{n}}, \text{ where } L_{\hat{\theta}} = \sum_{x \in X} \delta_x |\hat{\varrho}(x),$$

and * denotes convolution of laws.

It is well-known by Mecke's characterization of the Poisson process (see [17]) that P_y is characterized as the unique solution Q of the equation

(7.5)
$$C_Q(h) = \sum_{x \in X} \sum_{\gamma \in M} h(x, \gamma + \delta_x) \varrho(x) Q(d\gamma), h \in F_+.$$

To say it in another way. Q is the unique solution of the equation $C_Q(x, \gamma) = \varrho(x)Q(\gamma - \delta_x)$, $x \in X$, $\gamma \in M^-(X)$, $\gamma(x) \ge 1$. Another very useful view to equation (7.6) is

$$(7.6) \mathcal{C}_{\mathbf{Q}} = \mathcal{C}_{L_o} \star \mathbf{Q}.$$

(Note that the operation * differs from the convolution operation *.) To summarize: The first part of Theorem 6.1 implies Corollary 7.1. Let w be a statistical operator on \mathbb{C}^X with spectral law ϱ and z > 0 a parameter. Then $M_{\mathbb{R}^n}$ is a solution W of the equation $\mathfrak{C}_W = \mathfrak{C}_{L:n} * \kappa_W$.

This result is a version of Lemma 4.12 of Liebscher [16].

7.3. Bose-Einstein statistical operators with random particle number. We consider the Bose-Einstein statistical operator on the Fock space \mathbb{H}_+ with one-particle statistical operator w. It is clear that \mathbb{E}_w is symmetric and thereby also BE-symmetric. By the results obtained in B.4. \mathbb{E}_w is given by the following direct sum

(7.7)
$$\mathbb{E}_{w} = \frac{1}{\Xi_{w}^{+}} \sum_{m=0}^{\infty} tr(\Pi_{w}^{(m)} \mathbb{M}_{w}^{m}) \cdot \sum_{\gamma \in \mathcal{M}_{-}(X)} \mathsf{E}_{\varrho}^{m}(\gamma) \cdot Q_{e_{+}(\gamma)}^{+,m}.$$

Here we denote now the dependence on the particle number m in $Q_{c_{+}(\gamma)}^{+,m}$

Example 7.1. Consider a statistical operator w with ϱ being the uniform distribution on X, i.e. $\varrho \equiv \frac{1}{d}$. Recall that $d \geq 2$. In this case

$$tr(H_+^{(m)}\mathbb{M}_w^m) = \left(\begin{array}{c} d+m-1 \\ m \end{array} \right) \cdot \frac{1}{d^m};$$

and $\Xi^+ = \Xi^+_w(d) = \frac{1}{2}$. Thus the particle number distribution is given by the following negative binomial distribution

$$p_d^+(m) = \begin{pmatrix} d+m-1 \\ m \end{pmatrix} \cdot \left(1-\frac{1}{d}\right)^d \cdot \frac{1}{d^m}.$$

We want to calculate the Campbell measure C_{E_w} . Thus we first calculate its law: formulas (6.1) and (6.2) immediately imply that

(7.9)
$$\kappa_{\mathbb{E}_{\omega}} = \mathsf{E}_{\varrho} := \frac{1}{\Xi_{w}^{+}(d)} \sum_{m=0}^{\infty} tr(\Pi_{+}^{(m)} \mathbb{M}_{w}^{m}) \cdot \mathsf{E}_{\varrho}^{m}.$$

This point process is called here the *Bose-Einstein process* and denoted by \mathbb{E}_{ϱ} . This enables us to represent $\mathbb{E}_{u^{\ell}}$ as

$$\mathbb{E}_{w} = \sum_{\boldsymbol{\gamma} \in \mathcal{M}^{-}(X)} \mathsf{E}_{\varrho}(\boldsymbol{\gamma}) \cdot Q_{e_{+}(\boldsymbol{\gamma})}^{+}.$$

The Campbell measure of the Bose-Einstein statistical operator \mathbb{E}_w is given by the usual Campbell measure of the Bose-Einstein process. Moreover, \mathbb{E}_w is completely determined by the Campbell measure of its law \mathbb{E}_{ϱ} . So we have to study the Campbell measure $\mathfrak{C}_{\mathbb{E}_{\varrho}}$ which will be done in the Bab.

7.4. Fermi-Dirac statistical operators with random particle number. Consider now the Fermi-Dirac statistical operator on \mathbb{H}_{-} with one-particle statistical operator w. Analogously to the case of the Bose-Einstein operator it is FD-symmetric and can be represented as

$$\mathbb{D}_{w} = \frac{1}{\Xi_{w}^{-}} \sum_{m=0}^{\infty} \operatorname{tr}(\Pi_{-}^{(m)} \mathbb{M}_{w}^{m}) \cdot \sum_{\gamma \in \mathcal{M}_{m}(X)} \mathbb{D}_{v}^{m}(\gamma) \cdot Q_{\epsilon_{-}(\gamma)}^{-m}.$$

Example 7.2. Consider a statistical operator w with ϱ being the uniform distribution on X, i.e. $\varrho = \frac{1}{d}$ with $d \geq 2$. Then

$$\operatorname{tr}(\boldsymbol{\varPi}_{-}^{(m)}\mathbb{M}_{\boldsymbol{w}}^{m}) = \left(\begin{array}{c} \boldsymbol{d} \\ \boldsymbol{m} \end{array}\right) \cdot \frac{1}{\boldsymbol{d}^{m}};$$

and $\Xi^- = \Xi^-(d) = (1 + \frac{1}{d})^d$. Thus the particle number distribution is given by the following binomial distribution

$$p_{d}^{-}(m) = \begin{pmatrix} d \\ m \end{pmatrix} \cdot \left(\frac{1}{d+1}\right)^{m} \cdot \left(1 - \frac{1}{d+1}\right)^{d-m}.$$

Observe here the symmetry between Bose-Einstein and Fermi-Dirac statistical operators:

$$\Xi_w^-(d) = \Xi^+(-d).$$

We want to calculate its Campbell measure $C_{\mathfrak{D}_{\omega}}$. Again we calculate first its law: This is given by

(7.12)
$$\kappa_{\mathbb{D}_{w}} = \mathbb{D}_{\varrho} := \frac{1}{\Xi^{-}} \sum_{m=0}^{\infty} \operatorname{tr}(H_{-}^{(m)} \mathbb{M}_{m}^{m}) \cdot \mathbb{D}_{m}^{m}.$$

This point process is called the *Fermi-Dirac process* and is denoted by D_{ϱ} . Again we have a representation of the form

$$\mathbb{D}_{w} = \sum_{\gamma \in \mathcal{M}(X)} \mathsf{D}_{\varrho}(\gamma) \cdot Q_{e_{-}(\gamma)}^{-}.$$

Now we have the problem to study $\mathcal{C}_{D_{\varrho}}$ and to analyze D_{ϱ} . This problem will be solved in Bas by using again the method of the Campbell measure.

8. CHARACTERIZATIONS OF BOSE-EINSTEIN AND FERMI-DIRAC PROCESSES

The question is, what are the properties of the Boson resp. Fermion point processes. The answer is given by means of the method of the Campbell measure. For this aim we derive integration-by-parts formulas for E_{θ} resp. D_{θ} in terms of its Campbell measures. The arguments are only sketched. For the details we refer to [15, 20, 21, 25].

8.1. Bosons. Recall that the law ϱ on X is not a Dirac measure. Recall that for a given $\mu \in \mathcal{M}$ (X)

$$\mathsf{E}_{\nu}(\mu) = \frac{1}{\Xi_{i\nu}} \prod_{a \in X} \varrho(a)^{\mu(a)}.$$

If $\mu(X) = m$, this can be written as

$$\mathsf{E}_{\varrho}(\mu) = \frac{1}{\Xi_{\varrho}^{+}} \frac{1}{\binom{m}{\mu}} \mathsf{P}_{\varrho}^{m}(\mu).$$

In terms of the Poisson process in X with intensity measure ϱ , which is defined by

$$\mathsf{P}_{\varrho}(\mu) = e^{-\varrho(X)} \frac{\varrho(X)^m}{m!} \mathsf{P}_{\varrho}^m(\mu),$$

we obtain a representation of E₀ in terms of P₀:

(8.1)
$$\mathsf{E}_{\varrho}(\mu) = \frac{1}{\Xi_{\mathfrak{q}}^{+}} \frac{1}{\binom{m}{\mu}} \frac{m}{\varrho(X)^{m}} c^{\varrho(X)} \mathsf{P}_{\varrho}(\mu).$$

Now we start to calculate the Campbell measure of E_{ρ} , i.e.

$$\mathcal{C}_{\mathsf{E}_{\varrho}}(a,\mu) = \mu(a)\mathsf{E}_{\varrho}(\mu - \delta_a), \quad \mu(a) \ge 1$$

Using representation (8.1) in combination with Mecke's characterization (7.5) of the latter yields a recurrence which immediately leads to

Lemma 8.1. For $(a, \mu) \in C = \{(a, \mu) : \mu(a) \ge 1\}$

(8.2)
$$\mathbb{C}_{\mathsf{E}\varrho}(a,\mu) = \sum_{j=1}^{\mu(a)} \varrho(a)^j \cdot \mathsf{E}_\varrho(\mu - j\delta_a).$$

Observe that (8.2) is an equation for E_{ϱ} . To solve this equation we look at it in the following way:

Proposition 8.1. For any $h \in F_+$

(8.3)
$$\mathcal{C}_{\mathsf{E}_{\varrho}}(h) = \sum_{a \in X} \sum_{\gamma \in \mathcal{M}} \sum_{(X)} \sum_{j \geq 1} h(a, \gamma + j\delta_a) \varrho(a)^j \lambda(a) \mathsf{E}_{\varrho}(\gamma).$$

Here λ denotes the counting measure on X.

Equation (8.3) has the same structure as equation (7.6):

$$\left(\Sigma_{L_{v}^{+}}\right)$$
 $\mathcal{C}_{\mathsf{E}_{v}}=\mathcal{C}_{L^{+}}\mathsf{E}_{v}$

where the operation \star is a version of a convolution operation defined by the right hand side of (8.3); and L_0^+ is given by the following positive measure on $\mathcal{M}_f^-(X)$.

$$L_{\rho}^{+}(\varphi) = \sum_{j \geq 1} \sum_{\alpha \in X} \frac{1}{j} \varphi(j\delta_{\alpha}) \varrho(u)^{j}, \quad \varphi \in F_{+}.$$

This implies that E_{ρ} is the so called random KMM measure in X for L_{ρ}^{+} in the sense of [21].

As Mathias Raffer [25] has shown in full generality E_ϱ then coincides with the Polya sum process $\mathsf{S}_{\varrho,\lambda}$ for (ϱ,λ) . This process is by definition a Papangelou process with the kernel π^+ defined by

(8.4)
$$\pi^{+}(\mu, a) = \varrho(a) \cdot (\lambda(a) + \mu(a)), \quad a \in X, \mu \in \mathcal{M}(X).$$

And this means that $S_{\varrho,\lambda}$ is the unique solution S of the following integration by parts formula:

(8.5)
$$C_{S}(h) = \sum_{\mu} \sum_{a} h(a, \mu + \delta_{a}) \pi^{+}(\mu, a) S(\mu), \qquad h \in F_{+}.$$

This process has been called in [20] the *Pólya sum process for the parameters* (ρ, λ) . Thus we see that the characteristic properties of the Bose-Einstein process are twofold: It is a KMM process as well as a Pólya sum process.

The argument for the equality of E_{θ} and $\mathsf{S} = \mathsf{S}_{\theta,\lambda}$ is as follows: If one iterates the last equation (8.5) one obtains for any $N \in \mathbb{N}$

$$\begin{split} \mathcal{C}_{\mathsf{S}}(h) &= \sum_{\mu} \sum_{a} h(a, \mu + \delta_{a}) \varrho(a) (1 + \mu(a)) \mathsf{S}(\mu) \\ &= \sum_{j=1}^{N} \sum_{\mu} \sum_{a} \varrho(a)^{j} h(a, \mu + j\delta_{a}) \mathsf{S}(\mu) + \\ &+ \sum_{\mu} \sum_{a} \varrho(a)^{N} h(a, \mu + N\delta_{a}) \mu(a) \mathsf{S}(\mu) \\ &\longrightarrow_{N \to +\infty} \sum_{j \geq 1} \sum_{\mu} \sum_{a} \varrho(a)^{j} h(a, \mu + j\delta_{a}) \mathsf{S}(\mu) \end{split}$$

Here we used again that ϱ is not a Dirac measure and also that S is of first order. This shows that S solves equation (8.3) or equivalently (Σ_{L^+}) . One can show that this equation has only one solution. (Cf. [21]) To summarize we obtained the

Proposition 8.2. Given a probability ϱ on X which is not a Dirac measure then the Bose-Einstein process E_ϱ coincides with the random KMM measure in X for L^+ as well as the Polya sum process $\mathsf{S}_{\varrho,\lambda}$ for the parameters (ϱ,λ) . Moreover, this process is infinitely divisible and uniquely determined as a solution of the integration-by-parts formula (8.3).

We know also from [20] that the property of E_ϱ being a Papangelou process for π^* allows to calculate explicitly its particle number distribution. In the case where ϱ is the uniform distribution on X this coincides with p_+^+ which we calculated above

by completely different quantum mechanical methods. This implies that the point process in this case is of first order, i.e. the mean particle number is finite. (All this can be found in [20].) This shows that \mathbb{E}_{ϱ} has all properties of an ideal gas.

Moreover, equation (Σ_{L^*}) implies that E_ϱ is a socalled permanental process. This means that its reduced density matrix has a permanental structure. A proof based on (Σ_{L^*}) can be found in [21, 15] and the references therein.

Finally, using the above developed method of the Campbell measure, in particular Theorem 4, we obtain immediately characterizations of the Bose-Einstein statistical operator for w: The fact that $\kappa_{\mathbb{E}_w} = \mathbb{E}_\varrho$ solves equation $\left(\Sigma_{L^*}\right)$ immediately implies

Theorem 8.1. Let w be a statistical operator on \mathbb{C}^X with spectral law ϱ which is not a Dirac measure. A symmetric statistical operator W on the Fock space \mathbb{H}_+ , admitting a spectral resolution with respect to \mathbb{Y}_+ , coincides with \mathbb{E}_w iff it is a solution of equation $\mathbb{C}_W = \mathbb{C}_{L^+} \star \kappa_W$.

Moreover, $\kappa_{\mathbb{E}_{\varphi}} = \mathbb{E}_{\varphi}$ being also a solution to equation (8.5), implies

Theorem 8.2. Under the assumptions of Theorem 8.1, W coincides with \mathbb{E}_w iff it is the solution of the equation

(8.6)
$$\mathcal{C}_{W} h = \sum_{(x,\gamma)} h(x,\gamma+\delta_{x}) \pi^{+}(\gamma,x) \kappa_{W}(\gamma), \quad h \in F_{+}.$$

Statistical operators W which solve equation (8.6) can be called Pólya sum statistical operators specified by π_+ .

8.2. Fermions. The Campbell measure of D_{ϱ} is concentrated on C and given there by

$$C_{D_a}(a, \mu) = \varrho(a) \cdot D_{\varrho}(\mu - \delta_a), \quad \mu(a) = 1.$$

This implies that D_{ρ} is a Papangelou process for the kernel

$$\pi^{-}(a,\mu) = o(a) \cdot \left(\lambda(a) - \mu(a)\right), \qquad \mu(a) \leq 1;$$

(and $\pi^- \equiv 0$ else.) Recall here that λ denotes the counting measure. In the terminology of [20], D_{ϱ} is a *Polya difference process for* (λ,ϱ) . As for Bosons the distribution of the particle number is explicitly known, and the process is of first order. Again D_{ϱ} is completely determined by its kernel π_- , D_{ϱ} is a simple process, i.e. concentrated on $\mathcal{M}^+(X)$, and thus respects Pauli's exclusion principle. Furthermore, D_{ϱ} has independent increments. Thus it has all properties of an ideal gas. (For more details we refer to [20].) We observe here that the same reasoning we did above for the Papangelon process \mathbb{E}_{ϱ} yields that

Proposition 8.3. The Papangelou process D_a is the unique solution of the following equation for simple point processes Q.

(8.7)
$$C_{Q}(h) = \sum_{j=1}^{+\infty} (-1)^{j-1} \sum_{a,\mu} \varrho(a)^{j} h(a,\mu + j\delta_{a}) \ Q(\mu), \quad h \in F_{+}.$$

(The proof is exactly the same as above.) Again equation (8.10), which has D_{ϱ} as a unique solution, is of the form

$$\left(\Sigma_{L^{-}}\right) \qquad \qquad C_{\mathbf{Q}} = \mathbb{C}_{L^{-}} \star \mathbf{Q},$$

but now for the signed measure

$$L_{\varrho}^{-}(\varphi) = \sum_{j\geq 1} \sum_{a\in X} \frac{(-1)^{j-1}}{j} \varphi(j\delta_a) \varrho(a)^j, \quad \varphi \in F_+.$$

In this case one can show (see [21, 15]) that (Σ_L) implies that D_{ϱ} is a so called determinantal process.

As above for Bosons we obtain a characterization of symmetric statistical operators for Fermions: A symmetric statistical operator W, admitting a spectral resolution with respect to Y_- , coincides with D_w iff it is the unique solution of the equation $\mathcal{C}_W = \mathcal{C}_{L^-} \star \kappa_W$; or equivalently, iff it is the solution of the equation

$$C_W h = \sum_{(x,\gamma)} h(\gamma + \delta_x) \pi_-(\gamma, x) \kappa_W(\gamma), \quad h \in F_+.$$

Acknowledgement. We are grateful to Prof. Freudenberg and Dr. Rafler for several illuminating discussions. The second author is grateful for clarifying written discussions with Prof. Fichtner (Jena University) as well as to Prof. Goldin (Rutgers University). Comparing his approach (cf. [14]) to the one proposed in this paper, Prof. Goldin commented in a letter: "The general idea appears quite consistent with the approach I've taken in my work, but I don't think it is specifically contained in it. In particular, I've not explicitly made use of a basis in Hilbert space in order to construct the configuration space: rather. I've generally made use of configuration space and a group action on configuration space in order to construct various unitary group representations in Hilbert space." Finally, we thank the referee for several written discussions which led to a much better presentation of this work.

Список литературы

- [1] W. Arveson, An invitation to C*-algebras, Springer, New-York (1976).
- [2] A. Bach, Indistinguishable Classical Particles. Lecture notes in physics, Springer (1997)
- [3] F. A. Berezin, The Method of Second Quantization, New York, Academic Press (1966).
- [4] H. Boerner, Darstellungen von Gruppen, Springer (1955).

- [5] O. Bratteli, D. W. Robinson, Operator Algebras and Quantum Statistical Mechanics , Vol. 2 (2nd ed.), Springer (1997).
- [6] G. Cassinelli, N. Zanghi, "Conditional probabilities in quantum theory I conditioning with respect to a single event". Nuovo cimento, 73, 237 - 245 (1983).
- [7] J. M. Chaiken, "Finite-particle representations and states of the canonical commutation relations", Ann. Phys. 42, 23 80 (1967).
- [8] J. M. Cook, "The mathematics of second quantization", Transactions AMS 74, 222 245 (1953).
- [9] E. B. Davies, Quantum Theory of Open Systems, Academic Press (1976).
- [10] K. H. Fichtner, W. Freudenberg, "Point processes and the position distribution of infinite Boson systems", J. Statist. Phys. 47, 959 - 978 (1987).
- [11] K. H. Fichtner, W. Freudenberg, "Characterizations of states of infinite Boson systems", Commun. Math. Phys. 137, 315 - 357 (1991).
- [12] K. H. Fichtner, G. Winkler, "Generalized Brownian motion, point processes and stochastic calculus for random fields", Math. Nachr. 161, 291 – 307 (1993).
- [13] V. Fock, "Konfigurationsraum und zweite Quantelung", Zeitschrift für Physik 75, 622 647 (1932).
- [14] G. A. Goldin, U. Moschella, T. Sakuraba, "Self-similar processes and infinite-dimensional configuration spaces". American Inst. Physics, Physics of atomic nuclei 68, 1615 - 1684 (2005).
- [15] K. Krickeberg, Point Processes. A Random Radon Measure Approach, Nachst Neuendorf: Walter Warmuth Verlag (2014).
- [16] V. Liebscher, "Using weights for the description of states of Boson systems", Commun. Stoch. Anal. 3, 175 - 195 (2009).
- [17] J. Mecke, "Stationare Maße auf lokal-kompakten Abelschen Gruppen", Z. Wahrscheinlichkeitstheorie und verw. Gebiete 9, 36 58 (1967).
- [18] J. E. Moyal, "Particle populations and number operators in Quantum Theory", Adv. Appl. Prob. 4, 39 – 80 (1972).
- [19] B. Nehring, Construction of Classical and Quantum Gases. The Method of Cluster Expansions, Nachst Neuendorf: Walter Warmuth Verlag (2013).
- [20] B. Nehring, H. Zessin, "The Papangelou process. A concept for Fermion and Boson Processes", Izvestia NAN Armenii. Matematika 46, 49 66 (2011).
- [21] B. Nehring, H. Zessin, "A representation of the moment measures of the general ideal Bose gas". Math. Nachr. 285, 878 – 888 (2012).
- [22] J. von Neumann, Mathematische Grundlagen der Quantenmechanik, Springer Verlag, Berlin (1968).
- [23] M. Ozawa, "Conditional probability and a posteriori states in quantum mechanics", Publ. RIMS Kyoto Univ. 21, 279 – 295 (1985).
- [24] K. R. Parthasarathy, An Introduction to Quantum Stochastic Calculus, Birkhauser, Basel (1992)
- [25] M. Raffer, Gaussian Loop- and Pólya Processes. A point process approach. Universitätsverlag Potsdam (2009).
- [26] M. Rafler, "The Pólya sum process: Limit theorems for conditioned random fields", J. Theoretical Probab. 26, 1097 – 1116 (2013).
- [27] M. Raffer, H. Zessin, "The logical postulates of Boge, Carnap and Johnson in the context of Papangelou processes", Jt Theoret. Probab. DOI 10.1007/s10959-014-0543-2 (2014).
- [28] I. E. Segal, "Mathematical characterization of the physical vacuum for a linear Bose-Einstein field", Illinois J. Math. 6, 500 - 523 (1962).
- [29] H. Tamura, K. R. Ito, "A canonical ensemble approach to the Fermion/Boson random point processes and its applications", Commun. Math. Phys. 263, 353 – 380 (2006).
- [30] H. Zessin, Classical Symmetric Point Processes, Lectures held at the Instituto de Cibernetica, Matematica y Fisica (ICIMAF), La Habana, Cuba (2010).

Поступила 6 мая 2015

Известия ИАН Армении. Математика, том 52, н. 1. 2017, стр. 26-37.

ТОНКИЕ СВОЙСТВА ФУНКЦИЙ ИЗ КЛАССОВ ХАЙЛАША-СОБОЛЕВА M^p_α ПРИ p>0, II. АППРОКСИМАЦИЯ ЛУЗИНА

С. А. БОНДАРЕВ, В. Г. КРОТОВ

Белорусский государственный университет E-mails: bsa0393@gmail.com, krotov@bsu.by

Анпотация. В работе изучается впороксимация Лучина функций из классов Хайлаша. Соболева $M^p_\sigma(X)$ при p>0. Доказано, что для $f\in M^p_\sigma(X)$ и любого $\varepsilon>0$ существуют открытое множество $O_\varepsilon\subset X$, мера которого меньше ε (в качестве меры можно взять соответствующие емкость или вместимость Хаусдорфа), и прибліжающая функция f_ε такие, что $f=f_\varepsilon$ на $X\setminus O_\varepsilon$. При этом пеправляющая функция f_ε является регулярной (принадлежит исходному пространству $M^p_\sigma(X)$ классу и является локально гельдеровской) и приближает исходную функцию в метрике пространства $M^p_\sigma(X)$.

MSC2010 number: 46E35, 43A85.

Ключевые слова: Метрическое пространство с мерой, условие удвоения, класс Соболева, аппроксимация Лузина, емкость, внешняя мера, мера и размерность Хаусдорфа.

1. Введение

Наша работа является испосредственным продолжением работы [1]. Мы используем результаты из [1] для изучения свойства анпроксимации в смысле Дузина для классов Хайлаша-Соболева M^p_α при p>0. При этом мы полностью придерживаемся обозначений и определений из [1].

Теорема Лузина утверждает, что любая измеримая на \mathbb{R}^n функция f обладает C-свойством — f является непрерывной, если пренебречь множеством сколь угодно малой меры. Точнее, для любой измеримой на \mathbb{R}^n функции f и любого $\varepsilon > 0$ существуют такие функции $\varphi \in C(\mathbb{R}^n)$ и открытое множество $O_\varepsilon \subset \mathbb{R}^n$, для которых

$$f(x) = \varphi(x)$$
 upa $x \in \mathbb{R}^n \setminus O_{\varepsilon}$, $\mu(O_{\varepsilon}) < \varepsilon$

 (μ) мера Лебета на \mathbb{R}^n).

Если функция f является более регулярной в том или ином смысле, то исправляющая функция φ может обладать дополнительными свойствами гладкости и анпроксимирующими свойствами.

Для классов Хайлаша. Соболева M^p_{α} при $p\geq 1$ такие вопросы исследовались в [2]. [5]. Мы распространим результаты этих работ на случай p>0.

2. Основная теорема

Приведем определемия, необходимые для формулировки основного результата. Пусть (X,d,μ) — метрическое пространство с регулярной борелевской мерой μ и метрикой $d,\ B(x,r)=\{y\in X: d(x,y)< r\}$ — шар с центром в точке $x\in X$, радиуса r>0.

Будем предполагать, что мера μ удовлетворяет условию удвосния с ноказателем $\gamma>0$, т.е. для некоторой постоянной a_μ выполнено неравенство

$$\mu(B(x,R)) \le a_{\mu} \left(\frac{R}{r}\right)^{\gamma} \mu(B(x,r)), \quad x \in X, \ 0 < r \le R.$$

Для шара $B \subset X$ обозначаем r_B и x_B соответственно его радиус и центр, кроме того, $\lambda B =$ шар, концентрический с B, радиуса λr_B . Кроме того, пусть

$$f_B = \oint_B f d\mu = \frac{1}{\mu(B)} \int_B f d\mu$$

Через с всюду обозначаем различные положительные постоянные, зависящие, возможно, от определенных параметров, но эти зависимости для нас несущественны. Кроме того, запись $A \leq B$ всегда будет означать, что $A \leq cB$.

Неотрицательную функцию ν , определенную на σ алгебре борелевских множеств из X, будем называть внешней мерой, если она монотонна и субаддитивна с пекоторой постоянной a_{ν} , то есть для любой последовательности борелевских множеств E_k выполнено неравенство

$$\nu\left(\bigcup_{k} E_{k}\right) \leq a_{\nu} \sum_{k} \nu(E_{k}).$$

Кроме того, внешнюю меру будем называть регулярной в нуле, если для любого множества $E \subset X$ с $\nu(E) = 0$ и для любого $\varepsilon > 0$ существует открытое множество $O \supset E$, для которого $\nu(O) < \varepsilon$.

Пусть задана возрастающая функция $h:(0,1]\to(0,1], h(+0)=0$. Такие функции будем называть измеряющими. Мы будем использовать следующее условие, связывающее исходную меру μ и внешнюю меру ν : существует такая постоящая c_{ν} , что выполнено неравенство

$$(2.1) \nu(B) \le c_{\nu} \frac{\mu(B)}{h(r_B)} \quad \text{для всех шаров} \quad B \subset X, \ r_B \le 1.$$

Напомним, что для измеряющей функции h и $0 < R \le 1$ классическая и модифицированная (h,R)-вместимость Хаусдорфа множества $E \in X$ вводятся как

$$\begin{split} H_R^h(E) &= \inf \left\{ \sum_{i=1}^\infty h(r_i) : E \subset \bigcup_{i=1}^\infty B(x_i, r_i), \quad r_i < R \right\}, \\ \mathcal{H}_R^h(E) &= \inf \left\{ \sum_{i=1}^\infty \frac{\mu(B(x_i, r_i))}{h(r_i)} : E \subset \bigcup_{i=1}^\infty B(x_i, r_i), \quad r_i < R \right\}. \end{split}$$

соответственно (точная нижняя грань берется по всевозможным покрытиям множества E счетными семействами шаров). Величины

$$H^h(E) = \lim_{R \to +0} H^h_R(E), \quad \mathcal{H}^h(E) = \lim_{R \to +0} \mathcal{H}^h_R(E)$$

называются классической и модифицированной h-мерой Хаусдорфа для E соответственно. Для $h(t)=t^{\alpha}$, $\alpha>0$, пишем H^{α} и \mathcal{H}^{α} вместо $H^{t^{\alpha}}$ и \mathcal{H}^{t} . В случае классических мер также можно определить размерность Хаусдорфа

$$\dim_{\mathrm{H}} E = \inf \left\{ s : H_1^s(E) = 0 \right\}.$$

Введем класс Хайлаша Соболева $M^p_{\alpha}(X), \ 0 0$, как множество

$$M^p_\alpha(X) = \{ f \in L^p(X) : D_\alpha[f] \cap L^p(X) \neq \emptyset \},$$

$$||f||_{M^p_{\alpha}(X)} = ||f||_{L^p(X)} + \inf\{||g||_{L^p(X)} : g \in D_{\alpha}[f] \cap L^p(X)\},$$

где через $D_{\alpha}[f]$ обозпачен класс всех неотрицательных μ -измеримых функций g, для каждой из которых существует такое множество $E\subset X$, $\mu(E)=0$, что

$$|f(x) - f(y)| \le [d(x,y)]^{\alpha} [g(x) + g(y)], \quad x, y \in X \setminus E.$$

Классы $M^p_{\alpha}(X)$ порождают емкости

$$\operatorname{Cap}_{\alpha,p}(E)=\inf\left\{\|f\|_{M^p(X)}^p:f\in M^p(X),f\geq 1$$
 в окрестности $E\subset X
ight\}.$

Наконец, для $\alpha>0$ определим классы Гельдера

$$H_{\alpha}(X) = \left\{ \phi \in C(X) : \|\phi\|_{H_{\alpha}(X)} = \sup_{x \neq y} \frac{|\phi(x) - \phi(y)|}{|d(x, y)|^{\alpha}} < +\infty \right\}.$$

Наш основной результат формулируется следующим образом.

Теорема 2.1. Пусть $0 < \beta \le \alpha \le 1$, $0 и задана функция <math>f \in M^p_\alpha(X)$. Пусть также задана внешняя мера ν , регулярная в нуле и удовлетворяющия условию (2.1) с функцией $h(t) = t^{(n-1)p}$. Тогда для любого $\varepsilon > 0$ существуют функция f_ε и открытое множество $O \subset X$, такие, что

1)
$$\nu(0) < \varepsilon$$
.

- 2) $f = \int_{\mathcal{C}} na X \setminus O$.
- 3) $f_{\varepsilon} \in M^p_{\alpha}(X)$ u $f_{\varepsilon} \in H_{\beta}(B)$ для любого шара $B \subset X$.
- 4) $||f f_c||_{M^p(X)} < \varepsilon$.

В качестве примеров внешних мер, удовлетворяющих условию теоремы, можно взять $\nu = \operatorname{Cap}_{(\alpha-\beta),p}$ и $\nu = \mathcal{H}^{(\alpha-\beta)p}_1$, а также $\nu = H_1^{\gamma-(\alpha-\beta)p}$ (при условии $\mu(X) < \infty$).

При $\beta=\alpha=1,\ p>1$ подобный результат был ранее получен П.Хайлашем в [2], где вместо 1) утверждалось, что $\mu(O)<\varepsilon$, а в 3) было $f_{\varepsilon}\in H_1(X)$. Случай $\beta<\alpha=1$ теоремы 2.1 существенно сложнее, он был изучен в работе [3] при p>1 и в [4] при p=1. В работе [5] теорема 2.1 была доказана для $p>1,\ 0<\alpha\le 1$.

Следствие 2.1. Пусть $0 < \beta \le \alpha \le 1$, $0 и задана функция <math>f \in M^p_\alpha(X)$. Тогда для любого $\varepsilon > 0$ существуют функция f_ε и открытое множество $O \subset X$, такие, что

- 1) $\operatorname{Cap}_{(\alpha-\beta),p}(O) < \varepsilon$, $H_1^{\gamma-(\alpha-\beta)p}(O) < \varepsilon$, $\mathcal{H}_1^{(\alpha-\beta)p}(O) < \varepsilon$
- 2) $f = \int_{\mathbb{R}} na X \setminus O$,
- 3) $f_{\varepsilon} \in M^p_{\alpha}(X)$ и $f_{\varepsilon} \in H_{\beta}(B)$ для любого шара $B \subset X$.
- 4) $||f f_{\varepsilon}||_{M_{\alpha}^{\rho}(X)} < \varepsilon$.

Во время подготовки нашей работы к печати появился препринт [6], в котором также доказано утверждение следствии 2.1 для модифицированной вместимости Хаусдорфа. Методы [6] отличны от наших.

Результаты нашей работы докладывались на семинаре "Функциональные пространства" Университета Фридриха Шиллера (Иена, Германия, 19 декабря 2014 г. и 3 декабря 2015 г.), на Международной конференции "Функциональные пространства и теория аппроксимации функций", посвященной 110-летию со двя рождения академика С.М. Никольского (Москва, 28 мая 2015 г., и на Международной конференции «Нагшопіс Analysis and Approximations, VI» (Цахкадзор, Армения, 13 сентября 2015 г.).

При доказа гельстве теоремы 2.1 мы следуем схеме работ [3, 5]. Работу с несуммируемыми функциями обеспечивают результаты работ [1] и [9]. По существу, задача разбивается на две части. С одной стороны нам пужны кналифицированные оценки массивности лебеговых множеств некоторых максимальных функций, а с другой — надо уметь продолжать функции с этих множеств, сохраняя определенные условия гладкости. Для оценки исключительных множеств будем

С. А. БОНДАРЕВ, В. Г. КРОТОВ

использовать результаты из работ [10] и [1]. При этом важно использование некоторого анцроксимирующего анцарата — при $p \geq 1$ эту роль выполняют средние Стеклова. для $p \in (0,1)$ эта роль передается наилучшим L^p -приближениям постоянными на шарах $B \subset X$ [1]. Для продолжения функций применяется аналог конструкции Уитни, предложенный в [3] при рассмотрении аналогичной задачи для частного случая $\alpha = 1$.

3. Вспомогательные утверждения

Для доказательства основной теоремы нам понадобится ряд результатов, большинство из которых известны при p>1.

Лемма 3.1 ([1], лемма 9). Пусть $E \subset X$. $0 < \alpha \le 1$, $\gamma > \alpha p$. Тогда.

1) емкость Сира р является внешней мерой и

$$\operatorname{Cap}_{\alpha,p}(E) = \inf \left\{ \operatorname{Cap}_{\alpha,p}(O) : E \subset O, O - omkpumo \right\}.$$

2)
$$\operatorname{Cap}_{\alpha,r}(B(x,r)) \leq r^{-\alpha p} \mu(B(x,r))$$
 due $x \in X$. $0 < r \leq 1$,

3) npu
$$0 < \beta \le \alpha$$
 us $\operatorname{Cap}_{\alpha,p}(E) = 0$ enedyem $\operatorname{Cap}_{\beta,p}(E) = 0$.

Пусть

(3.1)
$$A_{p}(f,B) = \inf_{I} \left(\oint_{B} |f(y) - I|^{p} d\mu(y) \right)^{1/p}, \ p > 0,$$

тогда существует число $I_B^{(p)}f$ [9. лемма 3]), реализующее точную инжиюю границу в (3.1).

Техническим средством для докалательства основной теоремы является максимальный оператор $\dot{A}_{\alpha,R}^{(p)}f.$ Определим его следующим образом

$$\mathcal{A}_{\alpha,R}^{(p)}f(x) = \sup_{B\ni x, r_B \leqslant R} r_B^{-\alpha} A_p(f,B),$$

При $R=\infty$ вместо $\mathcal{A}_{\alpha,\infty}^{(p)}f$ будем писать $\mathcal{A}_{\alpha}^{(p)}f$.

Лемма 3.2 ([10]). Пусть p>0. $0 \le \beta < \alpha$. а мера μ и внешняя мера ν соязоны условием (2.1) с функцией $h(t)=t^{(\alpha-\beta)\nu}$. Тогда для $f\in L^p_{loc}(X)$ справедливо перавенство

$$\int_0^\infty \lambda^{p-1} \nu \{ \mathcal{A}_{\beta}^{(p)} f > \lambda \} d\lambda \lesssim \| \mathcal{A}_{\alpha}^{(p)} f \|_{L^p(X)}^p.$$

Лемма 3.3 ([9], лемма 3). *Пусть* $f \in L^p(X)$, p > 0. $B_1.B_2 \subset X$ — шари, причем $r_{B_1} < r_{B_2}$ и $B_1 \subset B_2$. Тогда

$$|I_{B_1}^{(p)}f - I_{B_2}^{(p)}f| \lesssim A_p(f, B_1) + \left(\frac{r_{B_2}}{r_{B_1}}\right)^{\frac{\gamma}{p}} A_p(f, B_2).$$

Лемма 3.4. Пусть $\beta, p>0, \ f\in L^p(X)$ и точка $x\in X$ такова, что

(3.2)
$$f(x) = \lim_{r \to +0} I_{B(x,r)}^{(p)} f.$$

Тогда

$$|f(x) - I_{B(x,r)}^{(p)} f| \lesssim r^{\beta} \mathcal{A}_{\beta}^{(p)} f(x).$$

 ${\it Доказательство}.$ Пусть для краткости $I_k = I_{B(x,2^{-k}r)}^{(n)} f.$ Тогда

$$|f(x) - I_0| = \left| \lim_{n \to \infty} \sum_{k=0}^{n-1} [I_k - I_{k+1}] \right| \le \sum_{k=0}^{\infty} |I_k - I_{k+1}|.$$

Далее используем лемму 3.3 для оценки каждого слагаемого в сумме

$$\sum_{k=0}^{\infty} |I_k - I_{k+1}| \lesssim \sum_{k=0}^{\infty} A_p(f, B(x, 2^{-k}r)) \lesssim \mathcal{A}_{\beta}^{(p)} f(x) \sum_{k=0}^{\infty} (2^{-k}r)^{\beta}.$$

Утверждение доказано.

Лемма 3.5. [11. лемма 2.5] Если $f \in M^p_{\alpha}(X)$, $\phi \in H_{\alpha}(X)$ и ограничена, то $f\phi \in M^p_{\alpha}(X)$. Кроме того, если $E \subset X$ и $\phi(x) = 0$ при $x \in X \setminus E$, то для любой функции $g \in D_{\alpha}(f) \cap L^p$

$$(g \|\phi\|_{\infty} + |f| \cdot \|\phi\|_{H^{\alpha}(X)}) \chi_E \in D_{\alpha}(f\phi) \cap L^p$$

Основным техническим средством для построения разбиений единицы и продолжения функции с сохранением гладкости является конструкция, изложенная в следующих двух леммах.

Лемма 3.6. [3, лемма 5.7] Нусть $O \subset X$ — открытое множество, $O \neq X$, $\mu(O) < \infty$. Для заданного $C \geq 2$ обозначим $r(x) = \frac{\operatorname{dist}(x, X \setminus O)}{2C}$. Тогда существует $N \geq 1$ и последовательность $\{x_1\}$ точек из X такие, что

- 1) wapot $B(x_i, r_i/4)$ nonapno не пересекаются, $r_i = r(x_i)$,
- 2) $\bigcup_{i=1}^{\infty} B(x_i, r_i) = O$,
- 3) $B(x_i, Cr_i) \subset O$,
- 4) ecan $x \in B(x_i, Cr_i)$, mo $Cr_i \leq \text{dist}(x, X \setminus O) \leq 3Cr_i$,
- 5) для любого і существуєт такое $y_i \in X \setminus O$, что $d(x_i,y_i) < 3Cr_i$,
- 6) $\sum_{i=1}^{\infty} \chi_{B(x_i,Cr_i)} \leq N.$

Лемма 3.7. [7. лемма 2.16] Пусть $0 < \alpha \le 1$, O открытое множество. $\{B(x_i,r_i)\}$ покрытие O шарами из леммы 3.6 для C=5. Тогда существует такая последовательность функций ϕ_i что

- 1) supp $\phi_i \subset B(\boldsymbol{x}_t, 2r_t), 0 \leq \phi_t(x) \leq 1$.
- 2) $|\phi_i(x) \phi_i(y)| \le cr^{-\alpha} [d(x, y)]^{\alpha}$.
- 3) $\sum_{i=1}^{\infty} \varphi_i(x) = \chi_O(x).$

4. Доказательство основной теоремы

Сначала сделаем дополнительное предположение – для некоторого $x_0 \in X$

$$(4.1) supp f \subset B(x_0, 1) \equiv B_0$$

Доказательство утверждения 1). В силу [1, теорема 3] $\nu(\Lambda)=0$, где Λ — множество точек $x\in X$, в которых не выполнено (3.2). Поэтому для $\varepsilon>0$ существует такое открытое множество $L\supset\Lambda$, что $\nu(L)<\varepsilon$. Для $\lambda>0$ обозначим

$$E_{\lambda} = \left\{ x \in X : A^{(p)} f(x) > \lambda \right\}.$$

Положим $O=E_\lambda\cup L$ и нокажем, что при достаточно большом λ множество O обладает необходимыми свойствами. Легко видеть, что O открыто и $O\in 2B_0$.

Из леммы 3.2

$$\int_0^\infty \lambda^{p-1} \nu(E_\lambda) \ d\lambda \lesssim \|\mathcal{A}_{\alpha}^{(p)} f\|_{L^p(X)}^p < \infty.$$

откуда следует, что $\nu(E_\lambda) \to 0$ при $\lambda \to +\infty$, кроме того, очевидно, что $\mu(E_\lambda) \to 0$ при $\lambda \to +\infty$.

Таким образом, утверждение 1) теоремы выполнено. Дополнительно выберем $\lambda > 0$ настолько большим, чтобы

$$(4.2) \qquad \int_{\Omega} |f|^p d\mu + \int_{\Omega} |A_{\alpha}^{(p)} f|^p d\mu < \varepsilon$$

Доказательство утверждения 2). Пусть $\{B(x_i,r_i)\}$ — покрытие множества O из леммы 3.6 для C=5. Тогда, применяя лемму 3.7, найдем набор функций $\{\phi_i\}_{i=1}^{\infty}$ таких, что

$$\operatorname{supp} \phi_i \subset B(x_i, 2r_i), \ 0 \leq \phi_i(x) \leq 1,$$
$$|\phi_i(x) - \phi_i(y)| \leq r^{-\alpha} \left[d(x, y) \right]^{\alpha}, \quad \sum_{i=1}^{\infty} \phi_i(x) = \chi_{\mathcal{O}}(x).$$

Определим функцию f_e равенством

$$f_{\varepsilon}(x) = \begin{cases} f(x), & x \in X \setminus O, \\ \sum_{i=1}^{\infty} \phi_{i}(x) f_{B(x_{i}, 2x_{i})}^{(p)} f_{\varepsilon} & x \in O \end{cases}$$

П

Утверждение 2 теоремы следует непосредственно из этого определения.

Доказательство утверждения 3). Сперва проведем вспомогательное рассуждение. Пусть $x \in O$, тогда существует такая точка $x^* \in X \setminus O$, что $d(x, x^*) \le 2 \operatorname{dist}(x, X \setminus O)$. Поэтому

$$|f_\varepsilon(x^*) - f_\varepsilon(x)| = \left|\sum_{i=1}^\infty \phi_i(x) \left[f(x^*) - I_{B(x_i, 2r_i)}^{(p)} f \right] \right| \leq \sum_{i \in I_x} |f(x^*) - I_{B(x_i, 2r_i)}^{(p)} f|,$$

где $I_x = \{i : x \in \text{supp } \phi_i\}.$

Заметим, что и точке $x^* \in X \setminus O$ выполнено соотношение (3.2), и $B(x_1, 2r_i) \subset B(x^*, 40r_i)$ для любого $i \in I_x$. Поэтому в силу лемм 3.3 и 3.4

$$|f(x^*) - I_{B(x_*, 2r_*)}^{(p)} f| \le |f(x^*) - I_{B(x^*, 40r_*)}^{(p)} f| + |I_{B(x_*, 2r_*)}^{(p)} f - I_{B(x^*, 40r_*)}^{(p)} f| \lesssim r_i^{\beta} A_{\beta}^{(p)} f(x^*).$$

Так как в I_x не более N слагаемых и $\mathcal{A}_{\beta}^{(p)}f(x^*) \leq \lambda$ (так как $x^* \in X \setminus O$), то

$$(4.4) \quad |f_{\varepsilon}(x^*) - f_{\varepsilon}(x)| \lesssim \sum_{i \in I_{\varepsilon}} r_i^{\beta} \mathcal{A}_{\beta}^{(p)} f(x^*) \lesssim \left[d(x, x^*) \right]^{\beta} \mathcal{A}_{\varepsilon}^{(p)} f(x^*) \lesssim \lambda \left[d(x, x^*) \right]^{\beta}.$$

Дальнейшее доказательство того, что $f_e \in H^{\beta}(X)$, проводится точно так же, как и при p>1 в работе [5], с той лишь разницей, что операторы \mathcal{S}_{β} заменяются на операторы $\mathcal{A}_{\beta}^{(p)}$. Для полноты повторим это рассуждение здесь. Рассмотрим три возможных случаи расположении точек $x,y\in X$.

Случай 1. Пусть $x, y \in X \setminus O$. Запишем очевидное неравенство

$$\begin{split} |f_{\varepsilon}(x) - f_{\varepsilon}(y)| &\leq |f(y) - I_{B(y,d(x,y))}^{(p)} f| + |f(x) - I_{B(x,2d(x,y))}^{(p)} f| + \\ &+ |I_{B(y,d(x,y))}^{(p)} f - I_{B(x,2d(x,y))}^{(p)} f|. \end{split}$$

Из леммы 3.4 следует, что первые два слагаемых мажорируются величиной

$$c\left[d(x,y)\right]^{\beta}\left[A_{\beta}^{(p)}f(x)+A_{\beta}^{(p)}f(y)\right].$$

Третье слагаемое также оценивается сверху этой же величиной в силу деммы 3.3. Таким образом, для $x,y\in X\setminus O$ выполнено

$$(4.5) |f_{\varepsilon}(x) - f_{\varepsilon}(y)| \leq |d(x,y)|^{\beta} \left[\mathcal{A}_{\beta}^{(p)} f(x) + \mathcal{A}_{\beta}^{(p)} f(y) \right] \leq \lambda \left[d(x,y) \right]^{\beta}$$

Случай 2. Пусть $x,y \in O$. Введем обозначение

$$(4.6) d_0 = \max \left\{ \operatorname{dist} (x, X \setminus O), \operatorname{dist} (y, X \setminus O) \right\}.$$

Если $d(x,y)>d_0$, то подберем точки $x^*,y^*\in X\setminus O$ так, чтобы $d(x,\varepsilon^*)\leq 2\operatorname{dist}(x,X\setminus O)$ в $d(y,y^*)\leq 2\operatorname{dist}(y,X\setminus O)$. Запишем очевидное неравенство

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| \leq |f_{\varepsilon}(x) - f_{\varepsilon}(x^*)| + |f_{\varepsilon}(y) - f_{\varepsilon}(y^*)| + |f_{\varepsilon}(x^*) - f_{\varepsilon}(y^*)|$$

Оценим первые два слагаемых с помощью (4.4), а гретье, используя (4.5). Получим

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| \lesssim \lambda [d(x, y)]^{\theta}$$
.

Пусть теперь $d(x,y) \le d_0$. Как и прежде, выберем $x^* \in X \setminus O$ так, чтобы $d(x,x^*) < 2 \operatorname{dist}(x,X \setminus O)$.

Поэтому, учитывая пункт 4) леммы 3.6, получаем

$$\begin{split} |f_{\varepsilon}(x) - f_{\varepsilon}(y)| &= \left| \sum_{i=1}^{\infty} \left[\varphi_{i}(x) - \varphi_{i}(y) \right] \left[I_{B(x_{i}, 2r_{i})}^{(p)} f - f(x^{*}) \right] \right| \lesssim \\ &\lesssim \left[d(x, y) \right]^{\alpha} \sum_{I_{x} \cup I_{y}} \frac{1}{r_{i}^{\alpha}} |I_{B(x_{i}, 2r_{i})}^{(p)} f - f(x^{*})|. \end{split}$$

Заметим, что для $i\in I_x\cup I_y$ справедливо включение $B(x_i,2r_i)\subset B(x^*,100r_i),$ следовательно, в силу лемм 3.3 и 3.4

$$|f(x^*) - I_{B(x_1, 2r_1)}^{(p)} f| \le |f(x^*) - I_{B(x^*, 100r_1)}^{(p)} f| + |I_{B(x_1, 2r_1)}^{(p)} f - I_{B(x^*, 100r_1)}^{(p)} f| \lesssim r_i^{\beta} \mathcal{A}_{\beta}^{(p)} f(x^*).$$

Таким образом, мы приходим к неравенству

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| \lesssim [d(x,y)]^{\beta} \sum_{I_x \cup I_y} \frac{[d(x,y)]^{\alpha - \beta}}{r_i^{\alpha - \beta}} \mathcal{A}_{\beta}^{(p)} f(x^*) \lesssim \lambda \left[d(x,y)\right]^{\beta}.$$

Cлучай 3. Пусть $x\in O$, а $y\in X\setminus O$. Выберем $x^*\in X\setminus O$ так, чтобы $d(x,x^*)\le 2\operatorname{dist}(x,X\setminus O)$. Тогда из (4.4) и уже доказанного пункта 1 следует

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| \le |f_{\varepsilon}(x) - f_{\varepsilon}(x^*)| + |f_{\varepsilon}(y) - f_{\varepsilon}(x^*)| \le \lambda \left[d(x,y)\right]^{\beta}$$

Таким образом, показано, что $f_{\varepsilon} \in H_{\delta}(X)$, если поситель функции f сосредоточен в единичном шаре.

Осталось показать, что $f_\varepsilon\in M^p_\alpha(X)$. Докажем сначала, что $f_\varepsilon\in L^p(X)$. Для этого оценим сверху $|I^{(p)}_{B(x_+,2r_+)}f|$: из (3.1) получаем

$$|I_{B(r_i,2r_i)}^{(p)}f|^p \lesssim A_p^p(f,B(x_i,2r_i)) + \int_{B(r_i,2r_i)} |f|^p d\mu \lesssim \int_{B(x_i,2r_i)} |f|^p d\mu.$$

Используя это неравенство и (4.3), имеем

$$\int_{O} |f_{\varepsilon}|^{p} d\mu \lesssim \sum_{i=1}^{\infty} \int_{B(x_{i}, 2r_{i})} |I_{B(x_{i}, 2r_{i})}^{(p)} f|^{p} d\mu =$$

$$(4.7) = c \sum_{i=1}^{\infty} \mu(B(x_i, 2r_i)) |I_{B(x_i, 2r_i)}^{(p)} f|^p \lesssim \sum_{i=1}^{\infty} \int_{B(x_i, 2r_i)} |f|^p d\mu = c \int_{\mathcal{O}} |f|^p d\mu.$$

Так как $f_{arepsilon}=f$ на $X \setminus O,$ то доказано, что $f_{arepsilon} \in L^p(X).$

Чтобы доказать, что $D_{\alpha}(f_{\varepsilon})\cap L^{p}\neq\varnothing$, покажем, что для некоторой постоянной с будет выполнено $c\mathcal{A}_{\alpha}^{(p)}f\in D_{\alpha}(f_{\varepsilon})\cap L^{p}$. Снова рассмотрим три различных случая расположения точек x,y.

1. Пусть $x,y\in X\setminus O$. Тогда, так как $f_{\varepsilon}=f$ на $X\setminus O$, то

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| = |f(x) - f(y)| \lesssim [d(x,y)]^{\alpha} \left[\mathcal{A}_{\alpha}^{(p)} f(x) + \mathcal{A}_{\alpha}^{(p)} f(y) \right].$$

2. Пусть $x,y\in O$. Предположим сначала, что $d(x,y)\leq d_0$. Так как, в силу леммы 3.7 $\|\phi_i\|_{H^\alpha(X)}\leq cr_i^{-\alpha}$, то

$$|f_{arepsilon}(x) - f_{arepsilon}(y)| = \left| \sum_{i=1}^{\infty} \left[\phi_i(x) - \phi_i(y) \right] \left[I_{B(x_i, 2r_i)}^{(p)} f - f(x) \right] \right| \lesssim$$

$$\lesssim \sum_{I \in \mathcal{U}} \frac{\left[d(x, y) \right]^{\alpha}}{r_i^{\alpha}} |I_{B(x_i, 2r_i)}^{(p)} f - f(x)|.$$

Так как при $i \in I_x \cup I_y$ выполняется включение $B(x_i, 2r_i) \subset B(x, 100r_i)$, то имеет место оценка

$$|I_{B(x_{i},2r_{i})}^{(p)}f-f(x)|\leq |I_{B(x,100r_{i})}^{(p)}f-f(x)|+|I_{B(x_{i},2r_{i})}^{(p)}f-I_{B(x,100r_{i})}^{(p)}f|\lesssim r^{\alpha}A_{\alpha}^{(p)}f(x)$$
 (см. леммы 3.3 и 3.4). Подставляя полученную оценку в (4.8) и используя условие 6) леммы 3.6, получим

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| \leq [d(x,y)]^{\alpha} \mathcal{A}^{(p)} f(x).$$

Теперь рассмотрим случай $d(x,y) > d_0$. Тогда

(4.8)

$$\begin{split} |f_{\varepsilon}(x) - f_{\varepsilon}(y)| &\leq \sum_{i \in I_{x}} \left| \phi_{i}(x) [I_{B(x_{i}, 2r_{i})}^{(p)} f - f(x)] \right| + \\ &+ \sum_{i \in I_{y}} \left| \phi_{i}(y) [I_{B(x_{i}, 2r_{i})}^{(p)} f - f(y)] | + |f(x) - f(y)| \leq \\ &\leq \sum_{i \in I_{x}} r_{i}^{\alpha} \mathcal{A}_{\alpha}^{(p)} f(x) + \sum_{i \in I_{i}} r_{i}^{\alpha} \mathcal{A}_{\alpha}^{(p)} f(y) + [d(x, y)]^{\alpha} \left[\mathcal{A}_{\alpha}^{(p)} f(x) + \mathcal{A}_{\alpha}^{(p)} f(y) \right] \lesssim \\ &\leq \left[\operatorname{dist} (x, X \setminus O) \right]^{\alpha} \mathcal{A}_{\alpha}^{(p)} f(x) + \left[\operatorname{dist} (y, X \setminus O) \right]^{\alpha} \mathcal{A}_{\alpha}^{(p)} f(y) + \\ &+ \left[d(x, y) \right]^{\alpha} \left[\mathcal{A}_{\alpha}^{(p)} f(x) + \mathcal{A}_{\alpha}^{(p)} f(y) \right] \lesssim \left[d(x, y) \right]^{\alpha} \left[\mathcal{A}_{\alpha}^{(p)} f(x) + \mathcal{A}_{\alpha}^{(p)} f(y) \right] \end{split}$$

3. Пусть $x \in O, y \in X \setminus O$. В этом случае получаем

$$|f_{\varepsilon}(x) - f_{\varepsilon}(y)| \leq \sum_{i \in I_x} |f(y) - I_{B(x_i, 2r_i)}^{(p)} f| \lesssim$$

$$\lesssim |f(y) - f(x)| + \sum_{i \in I_x} |f(x) - I_{B(x_i, 2r_i)}^{(p)} f| \lesssim [d(x, y)]^{\alpha} [\mathcal{A}_{\alpha}^{(p)} f(x) + \mathcal{A}_{\alpha}^{(p)} f(y)].$$

Таким образом, с $\mathcal{A}_{\alpha}^{(p)}f\in D_{\alpha}(f_{\varepsilon})\cap L^{p}$.

Доказательство утверждения 4). В силу (4.3), (4.7) и (4.2) $||f - f_e||_{L^p} \leq \varepsilon$. Кроме гого, если $g \in D_\alpha(f) \cap L^p$, то $c[A_\alpha^{(p)}f]_{\lambda O} \in D_\alpha(f - f_e) \cap L^p$. Тогда

$$\|f-f_{\varepsilon}\|_{M_{\alpha}^{p}} \leq \|f-f_{\varepsilon}\|_{L^{p}} + \|c[A_{\alpha}^{(p)}f]\chi_{O}\|_{L^{p}} \lesssim \varepsilon.$$

Избавимся теперь от предположения (4.1). Это делается точно так же, как и в случае p>1 (см. [3] и [5]). Действительно, существует не более чем счетный набор точек $\{x_i\}$, такой, что

$$X \subset \bigcup_{i} B(x_i, 1/2), \quad \partial(x_i, 1/4) \cap B(x_j, 1/4) = \emptyset \quad (i \neq j),$$

для которого можно построп гь другое разбиение единицы (см. [3]) — набор функций $\{\varphi_i\}\subset H_{\alpha}(X)$ со свойствами

$$0 \le \varphi_i(x) \le 1$$
, $\sup \varphi_i \subset B(x_i, 1)$, $\|\varphi_i\|_{H_\alpha(X)} = c$, $\sum_i \varphi_i(x) = 1$.

В силу леммы 3.5 $f\varphi_i \in M^p_\alpha(X)$, а так как supp $f\varphi_i \subset B(x_i,1)$, то в силу доказанного существует набор функций $\{f_\varepsilon\}$, удовлетворяющий условиям

$$f^i_{arepsilon}\in M^p_{lpha}(X)\cap H_{eta}(X),\quad \mathrm{supp}\, f^i_{arepsilon}\subset B(x_i,2),\quad \|f^i_{arepsilon}-farphi_i\|$$

При этом также

$$\nu\left\{x\in X: f_{\varepsilon}^{i}(x)\neq f\varphi_{i}(x)\right\}<\varepsilon/2^{i}.$$

Легко проверяется, что функция $f_* = \sum f_*^*$ удовлетворяет всем необходимым условиям. Теорема 2.1 доказана.

Abstract. The present paper is devoted to the Lusin's approximation of functions from Hajlasz–Sobolev classes $M^p(X)$ for p>0. It is proved that for any $f\in M^p_\alpha(X)$ and any $\varepsilon>0$ there exist an open set $O_\varepsilon\subset X$ with measure less than ε (as a measure can be taken the corresponding capacity or Hausdorff content) and an approximating function f_ε such that $f=f_\varepsilon$ on $X\setminus O_\varepsilon$. Moreover, the correcting function f_ε is regular (that is, it belongs to the underlying space $M^p_\alpha(X)$ and it is a locally Holder function), and it approximates the original function in the metric of the space $M^p_\alpha(X)$.

Список литературы

^[1] С. А. Боидарев. В. Г. Кротов, "Тонкие свойства функций из классов Хайлаша Соболева M^p_α при p>0, І. Точки Лебега", Изв. НАН Армении. Математика 51, по. 6, 3 – 22 (2016).

P. Hajiasz, "Sobolev spaces on an arbitrary metric spaces". Potential Anal. 5, no. 4, 403 - 415 (1996).

^[3] P. Hajlasz, J. Kinnunen, "Holder quasicontinuity of Sobolev functions on metric spaces", Rev. Mat. Iberoam. 14, no. 3, 601 622 (1998).

^[4] J. Kinnunen, H. Tuominen. Pointwise behaviour of M^{1,1} Sobolev functions, Math. Zeit. 257, no. 3, 613 - 630 (2007);

- [5] В. Г. Кротов, М. А. Прохорович, "Аппроксимация Лузина функций из классов W_n^{α} на метрических пространствах с мерой", Известия вузов. Математика, но. 5, 55 66 (2008).
- [6] T. Heikkinen, H. Tuominen, "Approximation by Holder functions in Besov and Triebel-Lizorkin spaces", preprint 2015, http://arxiv.org/abs/1504.02585.
- [7] R.A. Macias, C. Segovia "A decomposition into atoms of distributions on spaces of homogeneous type", Advances in Mathematics, 33, 271 - 309 (1979).
- [8] В. Г. Кротов, С. А. Бондарев "Тонкие свойства функций из пространств Хайлаша-Соболева $W_{\sigma}, \, p>0$ " Межд. конф. по функциональным пространствам и теории приближения функций, посв. 110-летию со дня рождения академика С. М. Никольского (Москва, 25–29 мая 2015 г.).
- [9] В. Г. Кротов, А. И. Порабкович, "Оценки L^p -осцилляций функций при p>0", Матем. заметки, 97, но. 3, 407 420 (2015).
- [10] В. Г. Кротов, "Весовые L^p -неравенства для шарп-максямальных функций на метрических пространствах с мерой". Изв. НАН Армении. Математика, 41. по. 2, 25 42 (2006).
- [11] J. Kinnunen, V. Latvala, "Lebesgue points for Sobolev functions on metric spaces", Rev. Mat. Iberoam. 18, no. 3, 685 700 (2002).

Поступила 3 февраля 2016

Известия ИАН Арменци, Математика, том 52. н. 1. 2017, стр. 38-46.

HYPERSURFACES OF A FINSLER SPACE WITH A SPECIAL (α, β) -METRIC

V. K. CHAUBEY AND A. MISHRA

Buddha Institute of Technology, Gida. Gorakhpur, India St. Joseph's College for Women, Gorakhpur, India E-mails: vkchaubey@outlook.com, arunimat6oct@hotmail.com

Abstract. In the present paper we study the Finslerian hypersurfaces of a Finsler space with a special (α, β) metric, and examine the hypersurfaces of this special metric as a hyperplane of first, second and third kinds.

MSC2010 numbers: 53B40, 53C60.

Keywords: Finslerian hypersurface, (α, β) metric.

1. Introduction

We consider an n-dimensional Finsler space $F^n=(M^n,L)$, that is, a pair consisting of an n-dimensional differentiable manifold M^n equipped with a Fundamental function L. The concept of an (α,β) metric, denoted by $L(\alpha,\beta)$, was introduced by M. Matsumoto [5], and later on has been studied by many authors (see [1 - 5, 8 - 9] and references therein). Well-known examples of (α,β) metrics are the Rander's metric $(\alpha+\beta)$, the Kropina metric $\frac{\alpha}{n}$ and the generalized Kropina metric $\frac{\alpha^{m+1}}{\beta m}(m\neq 0,-1)$. Recall that a Finsler metric L(x,y) is called an (α,β) metric if L is a positively homogeneous function of α and β of degree one, where $\alpha^2=a_{ij}(x)y^iy^j$ is a Riemannian metric and $\beta=b_i(x)y^i$ is an 1-form on M^n .

We consider a special Finsler Space $F^n=\{M^n,L(\alpha,\beta)\}$ with the metric $L(\alpha,\beta)$ given by

(1.1)
$$L(\alpha, \beta) = \alpha + \beta + \frac{\alpha^2}{(\alpha - \beta)}.$$

Differentiating equation (2.1) partially with respect to α and β , we get

$$\begin{split} L_{\alpha} &= \frac{2\alpha^2 + \beta^2 - 4\alpha\beta}{(\alpha - \beta)^2}, \quad L_{\beta} &= \frac{2\alpha^2 + \beta^2 - 2\alpha\beta}{(\alpha - \beta)^2}, \\ L_{\alpha\alpha} &= \frac{-2\beta^2}{(\alpha - \beta)^3}, \quad L_{\alpha\beta} &= \frac{-2\alpha\beta}{(\alpha - \beta)^3}. \end{split}$$

where

$$L_{\alpha} = \frac{\partial L}{\partial \alpha}$$
, $L_{\alpha} = \frac{\partial L}{\partial \beta}$, $L_{\alpha\alpha} = \frac{\partial L\alpha}{\partial \alpha}$, $L_{\beta\beta} = \frac{\partial L\beta}{\partial \beta}$, $L_{\alpha\beta} = \frac{\partial L\alpha}{\partial \beta}$

In the Finsler space $F^n = \{M^n, L(\alpha, \beta)\}$ the normalized element of the support $t_i = \partial_i L$ and the angular metric tensor h_{ij} are given by the following formulas (see [5]):

$$l_{ij} = \alpha^{-1}L_{\alpha}Y_{i} + L_{\beta}b_{i},$$

$$h_{ij} = pa_{ij} + q_{0}b_{i}b_{j} + q_{-1}(b_{i}Y_{j} + b_{j}Y_{i}) + q_{-2}Y_{i}Y_{j},$$

where $Y_i = a_{ij}y^j$. For the fundamental function (2.1) the constants p, q_0 , q_{-1} and q_{-2} in the last equation are given by the following formulas:

(1.2)
$$p = LL_{\alpha}\alpha^{-1} = \frac{4\alpha^4 - \beta^4 - 8\alpha^3\beta + 4\alpha\beta^3}{\alpha(\alpha - \beta)^3}.$$

$$q_0 = LL_{\beta\beta} = \frac{4\alpha^4 - 2\alpha^2\beta^2}{(\alpha - \beta)^4}, \quad q_{-1} = LL_{\alpha\beta}\alpha^{-1} = \frac{2\beta^3 - 4\alpha^2\beta}{(\alpha - \beta)^4}.$$

$$q_{-2} = L\alpha^{-2}(L_{\alpha\alpha} - L_{\alpha}\alpha^{-1}) = \frac{-4\alpha^5 - 2\alpha^2\beta^3 + 8\alpha^4\beta + \alpha\beta^4 - \beta^5}{\alpha^3(\alpha - \beta)^4}.$$

The fundamental metric tensor $g_{ij} = \frac{1}{2}\partial_i\partial_j L^2$ for $L = L(\alpha, \beta)$ is given by the following formula (see [4, 5]):

$$(1.3) g_{ij} = pa_{ij} + p_0b_ib_j + p_{-1}(b_iY_j + b_jY_i) + p_{-2}Y_iY_j$$

where

(1.4)
$$p_0 = q_0 + L_{\beta}^2 = \frac{8\alpha^4 + \beta^4 + 6\alpha^2\beta^2 - 8\alpha^3\beta - 4\alpha\beta^3}{(\alpha - \beta)^4},$$

$$p_{-1} = q_{-1} + L^{-1}pL_{\beta} = \frac{2\alpha\beta^3 - 4\alpha^3\beta + (2\alpha^2 + \beta^2 - 2\alpha\beta)^2}{\alpha(\alpha - \beta)^4}.$$

$$p_{-2} = q_{-2} + p^2L^{-2} = \frac{2\beta^4 + 8\alpha^2\beta^2 - 6\alpha\beta^3 + \frac{\beta^5}{\alpha}}{\alpha^2(\alpha - \beta)^3}.$$

The reciprocal tensor g^{ij} of g_{ij} is given by the following formula (see [4, 5]):

(1.5)
$$g^{ij} = p^{-1}a^{ij} - s_0b^ib^j - s_{-1}(b^iy^j + b^jy^i) - s_{-2}y^iy^j,$$

where $b^i = a^{ij}b_i$, $b^2 = a_{ij}b^ib^j$, and

(1.6)
$$s_{0} = \frac{1}{\tau p} \{ pp_{0} + (p_{0}p_{-2} - p_{-1})\alpha^{2} \},$$

$$s_{-1} = \frac{1}{\tau p} \{ pp_{-1} + (p_{0}p_{-2} - p_{-1})\beta \},$$

$$s_{-2} = \frac{1}{\tau p} \{ pp_{-2} + (p_{0}p_{-2} - p_{-1})b^{2} \},$$

$$\tau = p(p + p_{0}b^{2} + p_{-1}\beta) + (p_{0}p_{-2} - p_{-1})(\alpha^{2}b^{2} - \beta^{2}).$$

The hv-torsion tensor $C_{ijk} = \frac{1}{2} \partial_k g_{ij}$ is given by formula (see [10]):

$$(1.7) 2pC_{ijk} = p_{-1}(h_{ij}m_k + h_{jk}m_i + h_{ki}m_j) + \gamma_1 m_i m_j m_k,$$

where

(1.8)
$$\gamma_1 = p \frac{\partial p_0}{\partial \beta} - 3p_{-1}q_0, \quad m_1 = b_1 - \alpha^{-2}\beta Y_1.$$

Here m_i is a non-vanishing covariant vector orthogonal to the element of support y'.

Let $\{r_k\}$ be the component of the Christoffel symbol of the associated Riemannian space R^n , and let ∇_k be the covariant derivative with respect to x^k relative to this Christoffel symbol. Define

$$(1.9) 2E_{ij} = b_{ij} + b_{ji} 2F_{ij} = b_{ij} - b_{ji},$$

where $b_{ij} = \nabla_j b_i$.

Let $C\Gamma = (\Gamma_{jk}^{\bullet i}, \Gamma_{0k}^{\bullet}, \Gamma_{jk}^{\bullet})$ be the Cartan connection of F^n . The difference tensor $D_{jk}^i = \Gamma_{-k}^{\bullet i} - \{ i_k \}$ of the special Finsler space F^n is given by

$$(1.10) D_{jk} = B^{i} E_{jk} + F_{k}^{i} B_{j} + F_{j}^{i} B_{k} + B_{j} b_{0k} + B_{k}^{i} b_{0j} - b_{0m} g^{im} B_{jk}$$

$$- C_{jm}^{i} A_{k}^{m} - C_{km}^{i} A_{j}^{m} + C_{jkm} A_{s}^{m} g^{is} + \lambda^{s} (C_{jm}^{i} C_{sk}^{m} + C_{km}^{i} C_{sj}^{m} - C_{sk}^{m} C_{sk}^{m})$$

$$(1.10) D_{jk}^{i} = B^{i} E_{jk} + F_{k}^{i} B_{j} + F_{j}^{i} B_{k} + B_{j}^{i} b_{0k} + B_{k}^{i} b_{0j} - b_{0m} g^{im} B_{jk}$$

$$- C_{jm}^{i} A_{k}^{m} - C_{km}^{i} A_{j}^{m} + C_{jkm} A_{s}^{m} g^{is} + \lambda^{s} (C_{jm}^{i} C_{sk}^{m} + C_{sk}^{i} C_{sk}^{i} + C_{sk}^{$$

where

(1.11)
$$B_{k} = p_{0}b_{k} + p_{-1}Y_{k}, \quad B^{i} = g^{ij}B_{j}, \quad F_{i}^{k} = g^{kj}F_{ji},$$

$$B_{ij} = \frac{1}{2}\{p_{-1}(a_{ij} - \alpha^{-2}Y_{i}Y_{j}) + \frac{\partial p_{0}}{\partial \beta}m_{i}m_{j}\}, \quad B_{i}^{k} = g^{kj}B_{ji},$$

$$A_{k}^{m} = B_{k}^{m}E_{00} + B^{m}E_{k0} + B_{k}F_{i}^{m} + B_{0}F_{k}^{m},$$

$$\lambda^{m} = B^{m}E_{00} + 2B_{0}F_{0}^{m}, \quad B_{0} = B_{i}y^{i},$$

and '0' denotes the contraction with y' except for the quantities p_0, q_0 and s_0 .

2. INDUCED CARTAN CONNECTION

Let F^{n-1} be a hypersurface of F^n given by the equation $x^i = x^i(u^\alpha)$, where $\alpha = 1, 2, 3...(n-1)$. The element of the support y^i of F^n is taken to be tangential to F^{n-1} , that is, it is given by formula (see [6]):

$$(2.1) y^i = B^i_\alpha(u)v^\alpha.$$

The metric tensor $g_{\alpha\beta}$ and the hv-tensor $C_{\alpha\beta\gamma}$ of F^{n-1} are given by

$$g_{\alpha\beta} = g_{ij}B^i_{\alpha}B^j_{\beta}, \quad C_{\alpha\beta\gamma} = C_{ijk}B^i_{\alpha}B^j_{\beta}B^j_{\gamma},$$

and at each point (u^{α}) of F^{n-1} , a unit normal vector $N^{i}(u,v)$ is defined by

$$g_{ij}\{x(u,v),y(u,v)\}B^i_{\alpha}N^j=0, \quad g_{ij}\{x(u,v),y(u,v)\}N^iN^j=1.$$

The angular metric tensor $h_{\alpha\beta}$ of the hypersurface is determined by formulas:

$$(2.2) h_{\alpha\beta} = h_{ij}B^i_{\alpha}B^j_{\beta}, h_{ij}B^i_{\alpha}N^j = 0, h_{ij}N^iN^j = 1.$$

The inverse (B^{α}, N_i) of (B^i, N^i) is given by

$$egin{aligned} B_{\mathbf{i}}^{lpha} &= g^{lphaeta}g_{\mathbf{i}j}B_{eta}^{j}, & B_{lpha}^{\imath}B_{eta}^{eta} &= \delta_{lpha}^{eta}, & B^{lpha}N^{i} &= 0, \\ N_{i} &= g_{ij}N^{j}, & B^{k} &= g^{kj}B_{ji}, & B_{lpha}^{\imath}B^{lpha} + N^{\imath}N_{j} &= \delta_{j}^{\imath}. \end{aligned}$$

The induced connection $IC\Gamma = (\Gamma_{\beta}^{*\alpha}, G_{\beta}^{\alpha}, C_{\beta\gamma}^{\alpha})$ of F^{n-1} from the Cartan's connection $C\Gamma = (\Gamma_{k}^{*i}, \Gamma_{0k}^{*i}, C_{k}^{*i})$ is given by formulas (see [6]):

$$\begin{split} \Gamma^{**}_{0r} &= B^{\alpha}_{i} (B^{i}_{\beta\gamma} + \Gamma^{*i}_{jk} B^{j}_{\beta} B^{k}_{\gamma}) + M^{\alpha}_{\beta} H_{\gamma}, \\ G^{\alpha}_{\beta} &= B^{\alpha}_{i} (B^{i}_{0\sigma} + \Gamma^{*i}_{0i} B^{j}_{\gamma}), \qquad C^{\alpha}_{\beta\gamma} = B^{\alpha}_{i} C^{i}_{jk} B^{j}_{\beta} B^{k}_{\gamma}, \end{split}$$

where

$$M_{\beta\gamma} = N_i C^i_{jk} B^j_{\beta} B^k_{\gamma}$$
, $M^{\alpha}_{\beta} = g^{\alpha\gamma} M_{\beta\gamma}$, $H_{\beta} = N_i (B^i_{\alpha\beta} + \Gamma^{*i}_{\alpha\beta} B^{\dagger}_{\beta})$.

and

$$B_{\beta\gamma}^i = \frac{\partial B_{\beta}^i}{\partial v^{\gamma}}, \quad B_{0\beta}^i = B_{\alpha\beta}^i v^{\alpha}.$$

The quantities $M_{\beta\gamma}$ and H_{β} are called the second fundamental v-tensor and the normal curvature vector, respectively (see [6]). The second fundamental h-tensor $H_{\beta\gamma}$ is defined as follows (see [6]):

(2.3)
$$H_{\beta\gamma} = N_i (B_{\beta\gamma}^i + \Gamma_{ik}^{*i} B_{\beta}^i B^k) + M_{\beta\gamma}^{*i}$$

where

$$(2.4) M_0 = N_i C_{jk}^i B_3^j N^k.$$

The relative h- and v-covariant derivatives of the projection factor B'_{α} with respect to $IC\Gamma$ are given by

$$B_{\alpha|\beta}^{i} = H_{\alpha\beta}N^{i}, \quad B_{\alpha}^{i}|_{\beta} = M_{\alpha\beta}N^{i}.$$

It easily follows form equation (3.3) that $H_{\beta\gamma}$ generally is not symmetric and satisfies the equation

$$(2.5) H_{\beta\gamma} - H_{\gamma\beta} = M_{\beta}H_{\gamma} - M_{\gamma}H_{\beta}.$$

implying that

(2.6)
$$H_{0\gamma} = H_{\gamma}, \quad H_{\gamma 0} = H_{\gamma} + M_{\gamma} H_{0}.$$

The following lemmas, due to Matsumoto [6], will be used in Section 4

Lemma 2.1. The normal curvature $H_0 = H_\beta v^\beta$ vanishes if and only if the normal curvature vector H_β vanishes.

Lemma 2.2. A hypersurface $F^{(n-1)}$ is a hyperplane of the first kind with respect to the connection CT if and only if $H_{\alpha} = 0$.

Lemma 2.3. A hypersurface $F^{(n-1)}$ is a hyperplane of the second kind with respect to the connection CT if and only if $H_{\alpha} = 0$ and $H_{\alpha\beta} = 0$.

Lemma 2.4. A hypersurface $F^{(n-1)}$ is a hyperplane of the third kind with respect to the connection CT if and only if $H_{\alpha} = 0$ and $H_{\alpha\beta} = M_{\alpha\beta} = 0$.

3. A HYPERSURFACE $F^{(n-1)}(c)$ OF A SPECIAL FINSLER SPACE

Let us consider a Finsler space with the metric $L = \alpha + \beta + \frac{\alpha^2}{\alpha - \beta}$, where the vector field $b_i(x) = \frac{\partial h}{\partial x^2}$ is a gradient of some scalar function b(x). Now we consider a hypersurface $F^{(n-1)}(c)$ given by the equation b(x) = c, where c is a constant (see [10]). From the parametric equation $x^i = x^i(u^\alpha)$ of $F^{n-1}(c)$ we get

$$\frac{\partial b(x)}{\partial x^i} = 0,$$

$$\frac{\partial b(x)}{\partial x^i} \frac{\partial x^i}{\partial u^{ci}} = 0,$$

$$b_i B_0^i = 0,$$

showing that $b_i(x)$ is a covariant component of a normal vector field of the hypersurface $F^{n-1}(c)$. Further, we have

(3.1)
$$b_{\alpha}B_{\alpha}^{i}=0 \quad and \quad b_{i}y^{i}=0, \quad that \ is, \quad \beta=0,$$

and the induced matric L(u, v) of $F^{n-1}(c)$ is given by

(3.2)
$$L(u,v) = a_{\alpha\beta}v^{\alpha}v^{\beta}, a_{\alpha\beta} = a_{ij}B^{i}_{\alpha}B^{j}_{\beta},$$

which is a Riemannian metric.

Taking $\beta = 0$ in the equations (2.2), (2.3) and (2.5) we get

(33)
$$p = 4, \quad q_0 = 4, \quad q_{-1} = 0, \quad q_{-2} = -4\alpha^{-2}.$$

$$p_0 = 8, \quad p_{-1} = 4\alpha^{-1}, \quad p_{-2} = 0, \quad \tau = 16(1 + b^2).$$

$$s_{-1} = \frac{1}{4(1 + b^2)}, \quad s_{-1} = \frac{1}{4\alpha(1 + b^2)}, \quad s_{-2} = \frac{-b^2}{4\alpha^2(1 + b^2)}.$$

From (2.4) we get

$$(3.4) g^{ij} = \frac{1}{4}a^{ij} - \frac{1}{4(1+b^2)}b^ib^j - \frac{1}{4\alpha(1+b^2)}(b^iy^j + b^jy^i) + \frac{b^j}{4\alpha^2(1+b^2)}y^iv^j$$

Thus, from (4.1) and (4.4), along $F^{n-1}(c)$ we obtain

$$g^{ij}h_ib_j=\frac{1}{1+1+1}$$

Therefore we have

(3.5)
$$b_1(x(h)) = \sqrt{\frac{b^2}{4(1+b^2)}} N_t, \quad b^2 = \alpha^{ij} b_i b_j.$$

where b is the length of the vector b.

Next. from (4.4) and (4.5) we get

(3.6)
$$b' = a^{12}b_j = \sqrt{\frac{4b^2(1+b^2)}{\{1+b^2(1-\alpha^2)\}^2}}N + \frac{\alpha b^2 y}{1+b^2(1-\alpha^2)}.$$

Thus, we have the following result.

Theorem 3.1. In a special Finsler hypersurface $F^{(n-1)}(c)$, the induced Riemannian metric is given by (4.2) and the scalar function b(x) is given by (4.5) and (4.6).

Now, observe that the angular metric tensor h_{ij} and the metric tensor g_{ij} of F^n are given by formulas:

(3.7)
$$h_{ij} = 4a_{ij} + 4b_ib_j - \frac{4}{\alpha^2}Y_iY_j \quad and \quad g_{ij} = 4a_{ij} + 8b_ib_j + \frac{4}{\alpha}(b_iY_j + b_jY_i)$$

From equations (4.1), (4.7) and (3.2) it follows that if $h_{\alpha\beta}^{(a)}$ denotes the angular metric tensor of the Riemannian $a_{ij}(x)$, then along $F_{(c)}^{n-1}$ we have $h_{\alpha\beta} = h_{+}^{(a)}$. Thus, along $F_{(c)}^{n-1}$ we have $\frac{\partial a_{ij}}{\partial x} = \frac{\partial a_{ij}}{\partial x}$ and hence from equation (2.6) we get

$$Y_1 = \frac{48}{2}, \qquad m_i = b_i.$$

Therefore, in the special Finsler hypersurface $F_{(e)}^{(n-1)}$, the hv-torsion tensor becomes

(3.8)
$$C_{ijk} = \frac{1}{2a} (h_{ij}b_k + h_{jk}b_i + h_{ki}b_j) + \frac{6}{2}b_ib_jb_k.$$

Next, it follows from (3.2), (3.3), (3.5), (4.1) and (4.8) that

(3.9)
$$M_{\alpha\beta} = \frac{1}{2\alpha} \sqrt{\frac{b^2}{4(1+b^2)}} h_{\alpha\beta}$$
 and $M_{\alpha} = 0$.

Therefore, it follows from equation (3.6) that $H_{\alpha\beta}$ is symmetric. Thus, we have the following result.

Theorem 3.2. The second fundamental v-tensor of the special Finsler hypersurface $F_{(c)}^{(n-1)}$ is given by (4.9) and the second fundamental h-tensor $H_{\alpha\beta}$ is symmetric.

Now, from (4.1) we have $b_i B_{\alpha}^i = 0$, and hence

$$b_{i|\beta}B^{i}_{\alpha}+b_{i}B^{i}_{\alpha|\beta}=0.$$

Therefore, using the equality $b_{i|\beta} = b_{i|j}B_{\beta}^{j} + b_{i|j}N^{j}H_{\beta}$, from (3.5) we obtain

$$(3.10) b_{i|j}B_{\alpha}^{i}B_{\beta}^{j} + b_{i|j}B_{\alpha}^{i}N^{j}H_{\beta} + b_{i}H_{\alpha\beta}N^{j} = 0.$$

Since $b_i|_j = -b_h C_{ij}^{\dagger}$, we get $b_i|_i B_{\alpha}^i N^j = 0$. Therefore, taking into account that $b_{i|j}$ is symmetric, from equation (4.10) we have

(3.11)
$$\sqrt{\frac{b^2}{4(1+b^2)}} H_{\alpha\beta} + b_{\beta\beta} B_{\alpha}^{\beta} B_{\beta}^{\beta} = 0.$$

Next, contracting (4.11) with and using (3.1), we get

$$\sqrt{\frac{b^2}{4(1 + b^2)}} H_\alpha + b_{i|j} B_\alpha^i y^j = 0$$

Again contracting by v^{α} the equation (4.12) and using (3.1), we have

$$\sqrt{\frac{b^2}{4(1+b^2)}}H_0 + b_{0,j}y^*y^j = 0.$$

It follows from Lemmas 3.1 and 3.2 that the hypersurface F is a hyperplane of first kind if and only if $H_0 = 0$. Thus, in view of (4.13), it is obvious that $F_{(c)}^{n-1}$ is a hyperplane of first kind if and only if $b_{i|j}y^iy^j = 0$. On the other hand, $b_{i|j}$ being the covariant derivative with respect to $C\Gamma$ of F^n is defined on y^i , but $b_{ij} = \nabla_i b_i$ is the covariant derivative with respect to Riemannian connection $\binom{r}{k}$ constructed from $a_{ij}(x)$. Hence b_{ij} does not depend on y^i .

Below we consider the difference $b_{i|j} - b_{ij}$, where $b_{ij} = \nabla_j b_i$. The difference tensor $D_{jk} = \Gamma_{jk}^{*i} - \{^i_{jk}\}$ is given by (2.10), and since b_i is a gradient vector, then from (2.9) we have $E_{ij} = b_{ij}$, $F_{ij} = 0$ and $F_i^i = 0$. Thus, (2.10) reduces to the following

$$D_{jk} = B'b_{jk} + B_{j}b_{0k} + B_{k}^{i}b_{0j} - b_{0m}g^{im}B_{jk} - C_{jm}^{i}A_{k}^{ii} - C_{km}^{i}A_{j}^{im} + C_{jkm}A_{s}^{im}g^{is} + \lambda \left(C_{km}C_{km}^{i}C_{sj}^{mi} - C_{jk}^{m}C_{ms}^{i}\right),$$

where

(3.15)
$$B_{i} = 8b_{i} + 4\alpha^{-1}Y_{i}, \quad B^{i} = (\frac{1}{1+b^{2}})b^{i} + \frac{1}{\alpha(1+b^{2})}b^{i},$$

$$\lambda^{m} = B^{m}b_{00}, \quad B_{ij} = \frac{2}{\alpha}(a_{ij} - \frac{Y_{i}Y_{j}}{\alpha^{2}}) + \frac{12}{\alpha}b_{i}b_{j},$$

$$B^{i}_{j} = \frac{1}{2\alpha}(b^{i}_{j} - \alpha^{-1}y^{i}Y_{j}) + \frac{5}{2\alpha(1+b^{2})}b^{i}b_{j} - \frac{(1+6b^{2})}{2\alpha^{2}(1+b^{2})}b_{j}Y^{i}, \quad A^{m}_{k} = B^{m}_{k}b_{00} + B^{m}b_{k0}.$$

In view of (4.3) and (4.4), the relation in (2.11) becomes to by virtue of (4.15) we have $B_0^i = 0$, $B_{i0} = 0$ which leads $A_0^m = B^m b_{00}$.

Now contracting (4.14) by y^k we get

$$D_{j0}^{\iota} = B^{\iota}b_{j0} + B_{j}^{\iota}b_{00} - B^{m}C_{\tau m}^{\iota}b_{00}$$

Again contracting the above equation with respect to y^3 we obtain

$$D_{00}^i = B^i b_{00} = \{(\frac{1}{1+b^i})b^i + \frac{1}{a(1+b^i)}y^i\}b_{00}$$

In view of (4.1), along $F_{(c)}^{(n-1)}$ we get

$$(3.16) \quad b_i D_{j0}^i = \frac{b^2}{(1+b^2)} b_{j0} + \frac{(1+6b^2)}{2\alpha(1+b^2)} b_j b_{00} + \frac{1}{(1+b^2)} b_i b^m C_{jm}^i b_{00}$$

Now we contract (4.16) by y^j to obtain

(3.17)
$$b_i D_{00} = \frac{1}{(1+b^2)} b_{00}.$$

From (3.3), (4.5), (4.6), (4.9) and $M_{\alpha} = 0$ we obtain

$$b_i b^m C^i_{jm} B^j_\alpha = b^2 M_\alpha = 0.$$

Thus, the relation $b_{i|j} = b_{ij} - b_r D_{ij}^r$ and the equations (4.16), (4.17) give

$$b_{i|i}y^iy^j = b_{00} - b_v D_{00}^i = \frac{1}{1+b^2}b_{00}$$

Consequently, the equations (4.12) and (4.13) can be written as follows:

(3.18)
$$\sqrt{\frac{b^2}{4(1+b^2)}} H_{\alpha} + \frac{1}{1+b^2} b_{00} B_{\alpha}^* = 0$$

$$\sqrt{\frac{b^2}{4(1+b^2)}} H_0 + \frac{1}{1+b^2} b_{00} = 0.$$

Thus, the condition $H_0 = 0$ is equivalent to $b_{00} = 0$. Using the fact that $\beta = b_i y^i = 0$ the condition $b_{00} = 0$ can be written as $b_{ij}y^iy^j = b_iy^ib_jy^j$ for some $c_j(x)$. Therefore, we can write

$$(3.19) 2b_{ij} = b_i c_j + b_j c_i.$$

Now from (4.1) and (4.19) we get

$$b_{00} = 0, \quad b_{ij} B^{i}_{\alpha} B^{j}_{\beta} = 0, \quad b_{ij} B^{i}_{\alpha} y^{j} = 0.$$

It follows from (4.18) that $H_{\alpha}=0$, and hence in view of (4.15) and (4.19) we get $b_{i0}b^i=\frac{1}{2}$, $\lambda^m=0$, $A^iB^j_{\beta}=0$ and $B_{ij}B^i_{\alpha}B^j_{\beta}=0$ and

Next, we use the equations (3.3), (4.4) (4.6), (4.9) and (4.14) to obtain

$$(3.20) b_r D_{ij}^r B_{\alpha}^{\delta} B_{\beta}^{\delta} = -\frac{c_0 b^2 (4 + 3b^2)}{16\alpha (1 + b^2)^2} h_{\alpha\beta}.$$

Thus, the equation (4.11) reduces to the following

(3 21)
$$\sqrt{\frac{b^2}{4(1+b^2)}}H_{\alpha\beta} + \frac{b^2(4+3b^2)}{16\alpha(1+b^2)^2}h_{\alpha\beta} = 0.$$

and hence the hypersurface F_{00}^{n-1} is umbilic. Thus, we have the following result.

Theorem 3.3. A necessary and sufficient condition for F to be a hyperplane of first kind is (4.19). In this case the second fundamental tensor of Final is proportional to its angular metric tensor.

Now, taking into account that by Lemma 3.3, $F_{(c)}^{(n-1)}$ is a hyperplane of second kind if and only if $H_{\alpha} = 0$ and $H_{\alpha\beta} = 0$, from (4.20) we get $c_0 = c_i(x)y^i = 0$. Therefore, there exists a function $\psi(x)$ such that $c_i(x) = \psi(x)b_i(x)$, and, in view of (4.19), we get $2b_{ij} = b_i(x)\psi(x)b_j(x) + b_j(x)\psi(x)b_i(x)$. The last equation can also be written as follows $b_{ij} = \psi(x)b_ib_j$. Thus, we have the following result

Theorem 3.4. A necessary and sufficient condition for a hypersurface $F_{(c)}^{(n-1)}$ to be a hyperplane of second kind is (4.21)

Putting together Lemma 3.4 and formula (4.9), we conclude that $F_{(e)}^{n-1}$ is not a hyperplane of third kind. Thus, we have the following result.

Theorem 3.5. The hypersurface $F_{(n)}^{(n-1)}$ is not a hyperplane of the third kind.

Список литературы

- [1] M. Hashiguchi, Y. Ichijyo, 'On some special (α, β)-metric', Rep. Fac. Sci. Kagasima Univ. (Math., Phys., chem.) 8, 39 - 46 (1975).
- [2] S. Kikuchi, "On the condition that a space with (α, β) be locally Minkowskian", Tensor, N. S. 33, 242 246 (1979).
- [3] M. Kitayama, "On Finslerian hypersurfaces given by 3- change", Balkan Journal of Geometry and Its Applications, 7-2, 49 55 (2002).
- [4] M. Matsumoto, "Foundations of Finsler geometry and special Finsler spaces", Kaiseisha Press, Otsu. 520, Japan (1986).
- [5] M. Matsumoto, "Theory of Finsler spaces with (α, β)-metric". Rep. on Math. Phys. 31, 43 83 (1992).
- [6] M. Matsumoto, "The induced and intrinsic Finsler connections of a hypersurface and Finslerian projective geometry', J. Math. Kyoto Univ , 25, 107 - 144 (1985).
- [7] G. Randers, "On the asymmetrical metric in the four-space of general relativity". Phys. Rev. 2(59), 195 - 199 (1941).
- [8] C. Shibata, "On Finsler spaces with (α, β)", J. Hokkaido Univ. of Education, 35, 1 16 (1984)
- [9] I. Y. Lee, H. Y. Park, Y. D. Lee, "On a hypersurface of a special Finsler spaces with a metric
- $(\alpha+\frac{1}{\alpha})$ ", Korean J. Math. Sciences, 8, 93 101 (2001). [10] U. P. Saidi and B. Kumari, "On a Matsumoto space", Indian J. pure appl. Math., 32, 521 531 (2001)

О ЛОКАЛЬНОЙ ЭКВИВАЛЕНТНОСТИ МАЖОРАНТЫ ЧАСТИЧНЫХ СУММ И ФУНКЦИИ ПЭЛИ ДЛЯ РЯДОВ ФРАНКЛИНА

Г. Г. ГЕВОРКЯН, К. А. КЕРЯН

Epeванский государственный университет¹ E-mails: ggg@ysu.um, karenkeryan@ysu.am

Анпотация. Доказывается, что мажоранта частичных сумм и функция Пали ряда Франклина имеют эквивалентные нормы в пространстве $L_p(I)$, p>0, если интервалы "пика" функций Франклина с ненудевыми коэффициентами лежат в I. Приводятся примеры рядов указывающие на существенность этого условия

MSC2010 number: 42C10: 46E30.

Ключевые слова: Система Франклина: безусловный базис; функция Пэли.

1. Введение

Для формулировки полученных результатов, напомним определение системы Франклина. Пусть $n=2^{\mu}+\nu$, где $\mu=0,1,2,...$ и $1\leq\nu\leq 2^{\mu}$. Обозначим

$$s_{n,i} = \left\{ \begin{array}{ll} \frac{1}{2^{\mu+1}}, & \text{для} & 0 \leq i \leq 2\nu, \\ \\ \frac{1-1}{2^{\mu}}, & \text{для} & 2\nu < i \leq n. \end{array} \right.$$

Через S_n обозначим пространство функций, непрерывных и кусочно линейных на [0;1] с узлами $\{s_{n,i}\}_{i=0}^n$, т.е. $f\in S_n$, если $f\in C[0;1]$ и линейная на каждом отрезке $[s_{n,i-1};s_n]$, i=1,2,...,n. Ясно, что $\dim S_n=n+1$ и множество $\{s_{n,i}\}_{i=0}^n$ получается добавлением точки $s_{n,2\nu-1}$ к множеству $[s_{n,i}]_{i=0}^n$. Ноэтому, существует единственная, с точностью до знака, функция $f_n\in S_n$, которая ортогональна S_{n-1} и $\|f_n\|_2=1$. Полаган $f_0(x)=1$. $f_1(x)=\sqrt{3}(2x-1), x\in [0;1]$, нолучим ортонормированную систему $\{f_n(x)\}_{n=0}^\infty$, которая эквивалентным образом определена в работе [1] и называется системой Франклина.

Для $n=2^{\mu}+\nu$, где $\mu=0,1,2,...$, и $1\leq\nu\leq2^{\mu}$, обозначим (см. (1.1)) $\{n\}:=[s_{n,2\nu-2},s_{n,2\nu}]$ и $[n]=\mu$. Отрезок $\{n\}$ иногда называют интервалом пика функции

¹Исследования выполнены при финансовой поддержке ГКИ МОН РА в рамках научного проекта 15Т 14006

 f_n в связи с тем, что функция f_n достигает своего наименьшего и наибольшего значений на этом отрезке. Число [n] называют рангом числа n и отрезка $\{n\}$.

Систематическое изучение системы Франклина началось с работ [2], [3], где в частности доказано, что если $f \in L_p[0;1], 1 , и <math>\sum_{n=0}^{\infty} a_n f_n(x)$ ее ряд Фурьс-Франклина, то

(1.2)
$$S^*(f,\cdot) \in L_p[0;1]$$
, rge $S^*(f,x) = \sup_{n} |S_n(f,x)|$ и $S_n(f,x) = \sum_{k=0}^n a_k f_k(x)$.

C. В. Бочкаревым [4] была доказана, что система Франклина является безусловным базисом в пространстве $L_p[0;1],\ 1< p<\infty$. Для этого он доказал, что оператор Пэли для системы Франклина имеет слабый тип (1,1), т.е. существует постоянная C>0, такая что если $f\in L[0;1]$ и $\sum_{n=0}^{\infty}a_nf_n(x)$ ее ряд Фурье-Франклина, то

(1.3)
$$\max\{x \in [0,1] : P(f,x) > \lambda\} \le \frac{C}{\lambda} \int_0^1 |f(x)| dx,$$

где
$$P(f,x) = \left\{ \sum_{n=0}^{\infty} a_n^2 f_n^2(x) \right\}^{1/2}$$

Так как P имеет сильный тип (2,2), т.е. $\|P(f,\cdot)\|_2 \leq C\|f\|_2$, из (1.3), в силу известной интерноляционной теоремы Марцинкевича (см. папр. [5] стр. 485), следует, что для всех $p \in (1,\infty)$ имеет место $\|P(f,\cdot)\|_p \leq C_p\|f\|_p$. Следовательно, с учетом (1.2). для любого p > 1 имеем

(1.4)
$$\int_0^1 \sup_n \left| \sum_{k=0}^n a_k f_k(x) \right|^p dx \sim_p \int_0^1 \left(\sum_{k=0}^\infty a_k^2 f_k^2(x) \right)^{p/2} dx,$$

где запись $a\sim_{\gamma}b$ означает, что существуют постояпные c_{γ} и C_{γ} , зависящие только от γ , такие что $c_{\gamma}\cdot a\leq b\leq C_{\gamma}\cdot a$.

Из результатов работ [6] [8] следует, что (1.4) верно также для $p \in (0,1]$

Напомним, что пространство $L_0(E)$ — метрическое пространство п.в. конечных и измеримых на E функций с метрикой, сходимость по которой совпадает со сходимостью по мере на мпожестве E.

В работе [9] доказан не только аналог соотношения (1.4) в случае p=0, но и ее локализация на множества положительной меры. А именно, доказана следующая (см. георемы 2.1, 2.2 и 2.3 работы [9])

Теорема 1.1. Для ряда $\sum\limits_{k=0}^{\infty}a_kf_k(x)$ следующие условия эконовлентны:

- (1) ряд почти всюду сходится на Е,
- (2) ряд по мере безусловно сходится на Е,

о локальной эквивалентности мажоранты частичных сумм ...

(3)
$$\sup_{n} |\sum_{k=0}^{n} a_k f_k(x)| < +\infty$$
 n.s. na E ,

(4)
$$\sum_{k=0}^{\infty} a_k^2 f_k^2(x) < +\infty$$
 n.s. nn E.

В настоящей работе мы докажем, что невозможно получить локализацию соотношения (1.4) даже на двоичных интервалах и получим такую локализацию при некотором дополнительном условии.

2. Основные результаты

Теорема 2.1. Для любого двоичного отрезка $I = \left[\frac{a^*}{2^k}, \frac{p+1}{2^{k-1}}\right]$, любого рядо $\sum a_n f_n(x)$ и любого p>0 имсет место

$$\left\|\sup_{N}\left|\sum_{\{n\}\subset I,n\leq N}a_{n}f_{n}(x)\right|\right\|_{L_{p}(I)}\sim_{p}\left\|\left\{\sum_{\{n\}\subset I}a_{n}^{2}f_{n}^{2}(x)\right\}^{\frac{1}{2}}\right\|_{L_{p}(I)}$$

Теорема 2.2. Для любого отрежа $I = \left[\frac{ar}{2^k}, \frac{n+1}{2^k}\right] \neq [0,1]$, любых p > 0 и C > 0 существуют ряды $\sum_{n=0}^{\infty} a_n f_n(\tau)$ и $\sum_{n=0}^{\infty} b_n f_n(\tau)$ такие, что

(2.1)
$$\left\| \sup_{N} \left| \sum_{n \leq N} a_n f_n(x) \right| \right\|_{L_p(I)} > C \cdot \left\| \left\{ \sum_{n=0}^{\infty} a_n^2 f_n^2(x) \right\}^{\frac{1}{2}} \right\|_{L_p(I)}$$

(2.2)
$$\left\| \left\{ \sum_{n=0}^{\infty} b_n^2 f_n^2(x) \right\}^{\frac{1}{2}} \right\|_{L_p(I)} > C \cdot \left\| \sup_{N} \left| \sum_{n \le N} b_n f_n(x) \right| \right\|_{L_p(I)}$$

Теорема 2.2 указывает на то, что условие $\{n\} \subset I$ в теореме 2.1 существенно. Мы убедимся, что для системы Хаара не верен аналог теоремы 2.2, а аналог теоремы 2.1 верен и без условия $\{n\} \subset I$. Однако в этом случае постоянные эквивалентности также зависят от I. А именно верна следующая

Теорема 2.3. Для любого двоичного отрезка $I = \left[\frac{w}{2^k}, \frac{w+1}{2^k}\right]$, любого p > 0 и любого ряда $\sum a_n \chi_n(x)$ имеет место

$$\left\| \sup_{N} \left| \sum_{n \leq N} a_n \chi_n(x) \right| \right\|_{L_p(I)} \sim_{p,I} \left\| \left\{ \sum_{n} a_n^2 \chi_n^2(x) \right\}^{\frac{1}{2}} \right\|_{L_p(I)}$$

Учитыная соотношение (1.4), для доказательства теоремы 2.1, достаточно доказать следующую лемму. **Лемма 2.1.** Для любого двоичного отрезка $I = \left[\frac{w}{2^k}, \frac{w-1}{2^k}\right]$, любого p > 0 и любого ряда $\sum a_n f_n(x)$ имеют место

$$(2.3) \qquad \int_0^1 \sup_N \left| \sum_{\{n\} \subseteq I, n \le N} a_n f_n(x) \right|^p dx \le C_p \cdot \int_I \sup_N \left| \sum_{\{n\} \subseteq I, n \le N} a_n f_n(x) \right|^p dx,$$

(2.4)
$$\int_{0}^{1} \left\{ \sum_{\{n\} \subset I} a_{n}^{2} f_{n}^{2}(x) \right\}^{\frac{1}{2}} dx \leq C_{p} \cdot \int_{0}^{1} \left\{ \sum_{\{n\} \subset I} a_{n}^{2} f_{n}^{2}(x) \right\}^{\frac{1}{2}} dx.$$

Для доказательства вышесформулированных утверждений напомним некоторые свойства системы Франклина.

В исследованиях рядов по системе Франклина важную роль играют так называемые экспоненциальные оценки, полученные 3. Чисельским [3]

$$C_1 \cdot 2^{rac{|n|}{2}} \cdot (2-\sqrt{3})^{|i-2\nu-1|} < (-1)^{i+1} f_n(s_{n,i}) < C_2 \cdot 2^{rac{|n|}{2}} \cdot (2-\sqrt{3})^{|i-2\nu-1|},$$
rge $C_1 = rac{2+\sqrt{3}}{3\sqrt{3}}$ u $C_2 = 4 \cdot \sqrt{3}(2+\sqrt{3}).$

Для доказательства леммы 2.1 удобнее использовать оценки, полученные в [10]. Их сформулируем в виде Предложений 2.1 и 2.2.

Предложение 2.1. Для любого $n=2^{\mu}+\nu$ имеют место следующие неравенства:

$$(2.5) \qquad \frac{1}{4}|f_n(s_{n,i+1})| < |f_n(s_{n,i})| < \frac{2}{7}|f_n(s_{n,i+1})| \quad \text{korda} \quad 1 \le i < 2\nu - 2,$$

$$(2.6) \qquad \frac{1}{4}|f_n(s_{n,i-1})| < |f_n(s_{n,i})| < \frac{2}{7}|f_n(s_{n,i-1})| \quad \text{koeda} \quad 2\nu < i < n.$$

$$(2.7) \ f_n(s_{n,1}) = -2f_n(s_{n,0}), \quad f_n(s_{n,n-1}) = -2f_n(s_{n,n}), \quad f_n(s_{n,1}) \cdot f_n(s_{n,n+1}) < 0.$$

Предложение 2.2. Для $n=2^k+\nu,\ c\ условисм\ 1<\nu<2^k,\ выполняются$

$$(2.8) \qquad \frac{97}{48} < \frac{|f_n(s_{n,2\nu-1})|}{|f_n(s_{n,2\nu})|} < \frac{95}{42} \qquad \frac{107}{66} < \frac{|f_n(s_{n,2\nu-1})|}{|f_n(s_{n,2\nu-2})|} < \frac{98}{60}.$$

Из Предложения 2.1 и липейности функции f_n на $[s_{n,i-1},s_{n,i}]$ простыми вычислениями легко выводится

Предложение 2.3: Для $n=2^k+\nu,\ c$ условием $1<\nu<2^k,$ любого p>0 выполняются

$$(2.9) \qquad \int_{-\pi/2}^{\pi/2} |f_n(x)|^p dx < \left(\frac{2}{7}\right)^p \cdot \int_{-\pi/2}^{\pi/2} |f_n(x)|^p dx, \quad \text{kolds} \quad i \le 2\nu - 2$$

74

$$\int_{-\pi}^{\pi} |f_n(x)|^p dx < \left(rac{2}{\pi}
ight)^{-1} \int_{-\pi}^{\pi} |f_n(x)|^p dx$$
, words $i \geq 2
u + 2$.

Предложение 2.4. Пусть $n_1=2^{\mu}+\nu_1$. $n_2=2^{\mu}+\nu_2$ и $\nu_1<\nu_2$. Тогда существуют числа α , β (зависящие от μ , ν_1 и ν_2), такие, что

$$f_{n_1}(x) = \alpha \cdot f_{n_2}(x)$$
, korda $x \leq s_{n_1,2\nu_1-2} = \frac{\nu_1 - 1}{2^{\mu}}$

u

$$f_{n_1}(x) = \beta \cdot f_{n_2}(x), \quad \text{korda} \quad x \ge s_{n_2,2\nu_2} = \frac{\nu_2}{2^p}.$$

Это Предложение впервые было применено в [8]. Из этого Предложения немедленно следует

Предложение 2.5. Для любых $a_n,\ n=2^\mu+\nu,\ 1\le\nu\le 2^\mu$ и любого $\nu_0,\ 1\le\nu_0\le 2^\mu$ имеют место

$$\sum_{\nu=1}^{\nu_0} a_{2^{\mu}+\nu} f_{2^{\mu}+\nu}(x) = \alpha \cdot f_{2^{\mu}+\nu_0}(x), \quad \text{norda} \quad x \geq \frac{\nu_0}{2^{\mu}}$$

21

$$\sum_{\nu=\nu_0}^{2^{\mu}} a_{2^{\nu}+\nu} f_{2^{\nu}+\nu}(x) = \beta \cdot f_{2^{\mu}+\nu_0}(x), \quad \text{koeda} \quad x \leq \frac{\nu_0-1}{2^{\mu}}.$$

где α и β некоторые числа зависящие от a_n

Условимся через C, C_1, C_{γ} , обозначать постоянные, зависящие только от своих индексов. Значения этих постоянных в разных формулах могут быть разными. Длину отрезка I обозначим через |I|.

Доказательство леммы 2.1. Сначала докажем соотношение (2.3). Вместо $\sum_{\{n\} \subset I} a_n f_n(x)$ будем писать $\sum_n a_n f_n(x)$, предполагая $a_n = 0$, когда $\{n\} \not\subset I$. Допустим

(2.10)
$$\int_{f} (S^{*}(x))^{p} dx = 1, \text{ rge } S^{*}(x) = \sup_{N} \sum_{n \leq N} a_{n} f_{n}(x) ,$$

и докажем, что

(2.11)
$$\int_{0}^{\frac{\pi}{2^{k}}} (S^{*}(x))^{p} dx \leq C.$$

Положим для $n_m = 2^m + w \cdot 2^{m-k} + 1$.

$$\sigma_m(x) = |a_{n_m} f_{n_m}(x)|, \quad \text{if} \quad S_m^{\star}(x) = \max_N \left| \sum_{[n] = m, n \leq N} a_n f_n(x) - a_{n_m} f_{n_m}(x) \right|.$$

Очевидио, что

$$(2.12)$$
 $\sigma_m(x) \le 2 \cdot S^*(x), \quad S^*_m(x) \le 2 \cdot S^*(x), \quad \text{когда} \quad x \in [0,1]$

И

$$S^*(x) \leq \sum_{m=1}^\infty \sigma_m(x) + \sum_{m=1}^\infty S_m(x) =: \Sigma_1(x) + \Sigma_2(x)$$
. когда $x \in [0,1]$.

Мы докажем, что

(2.13)
$$\int_{0}^{\frac{\pi}{2^{n}}} \Sigma_{i}^{p}(\pi) dx \leq C_{p}, \quad i = 1, 2.$$

Пусть I_m - слева первый интервал ранга m, который содержится в I. а I_m^* -его правал половина. Заметим, что отрезки I_m^* не пересекцются и их объединение есть отрезок I. Заметим также, что функции $\sigma_m(x)$ и $S_m^*(x)$, $m \geq k$. являются модулями линейных на отрезке I_m^* функций. Действительно, $\sigma_m(x)$ -модуль от линейной на I_m^* функции $\sigma_m f_{n_m}(x)$. А для $S_m^*(x)$, в силу Предложения 2.5 имсем, что для любого N существует σ_N , такое что

$$(2.14) \sum_{[n]=m,n\leq N} a_n f_n(x) - a_{n_m} f_{n_m}(x) = \alpha_N \cdot f_{n_m+1}(x) \text{ and } x \leq \frac{\infty}{2^k} + \frac{1}{2^m}$$

Отсюда имеем

$$(2.15) S_m^*(x) = |f_{n_m+1}(x)| \max_N |\alpha_N|, \text{ and } x \leq \frac{w}{2^k} + \frac{1}{2^m}$$

Через A_m и B_m обозначим интегральные средние на I_m^* функций $\sigma_m(x)$ и $S_m^*(x)$, соответственно, т.е.

$$(2.16) A_m = \frac{1}{|I_m^*|} \int_{I^*} \sigma_m(x) dx, \quad \text{if } B_m = \frac{1}{|I_m^*|} \int_{I_m^*} S_m^*(x) dx, \quad m \ge k.$$

Нз того, что $\sigma_m(x)$ и $S_m^*(x), \, m \geq k$, являются модулями линейных на отрезке I_m^* функций, получаются

(2.17)
$$A_m \sim \max \sigma_m(x) \quad \text{if } B_m \sim \max S_m^*(x).$$

Из (2.16). (2.12) и (2.10), с применением неравенства Гелдера, имеем

$$\sum_{m \ge k} A_m^p |I_m^*| = \sum_{m \ge k} |I_m^*| \cdot |I_m^*|^{-p} \left(\int_{I_m^*} \sigma_m(x) dx \right)^p \le$$

$$\sum_{m \ge k} |I_m^*|^{1-p} \int_{I_m^*} \sigma_m^p(x) dx \cdot |I_m^*|^{p-1} \le 2^p \sum_{m \ge k} \int_{I_m^*} (S^*(x))^p dx = 2^p$$

Аналогично получаем

(2.18)
$$\sum_{m \ge k} B_m^{\mu} |I_m^*| \le 2^{\mu}.$$

Сначала докажем неравенство (2.13) для i=2 и $p\le 1$. В этом случае, последовительно применяя (2.9), (2.12) и (2.10) получим

$$\int_{0}^{\frac{s}{p}} \Sigma_{2}^{p}(x) dx = \int_{0}^{\infty} \left(\sum_{m} S_{m}^{*}(x) \right)^{p} dx \le \int_{0}^{\infty} \sum_{m} \left(S_{m}^{*}(x) \right)^{p} dx \le C_{p} \cdot \int_{I} \left(S_{m}^{*}(x) \right)^{p} dx \le C_{p}.$$

Аналогично получается неравенство (2.13), когда i=1 и $p \le 1$.

Перейдем к получению оценок (2.13), когда p>1. Без ограничения общности можем считать, что суммы Σ_1 , i=1,2, копечны, т. е. $\Sigma_1=\sum_{m=k}^k S_m^*$.

Через I_m^j , $k \le m \le k_1$, обозначим двоичные отрезки ранга m+1, где правый конец отрезка I_m^1 совпадает с левым концом отрезка I, а для j>1 правый конец отрезка I_m^j совпадает с левым концом I_m^{j-1} .

Для $x \in [0, \frac{m}{2^k}]$ и $m \ge k$ обозначим

(2.19)
$$\varphi_m(x) := B_{m,j} := \frac{1}{|I_m^j|} \int_{I_m^j} S_m^*(t) dt$$
, koraa $x \in I_m^j$.

Поскольку $S_m^*(x)$ -модуль от линейной на I_m^* функции, то (здесь и далее полагается q=2/7) из (2.15), (2.17) и (2.9) имеем

$$(2.20)$$
 $S_m^*(x) \le C \varphi_m(x) \le C B_{m,j} \le C q^{j-1} B_{m,1} \le C q^{j-1} B_m$, когда $x \in I_m^j$.

Обозначим $J_m = I_m^1 \backslash I_{m+1}$, когда $m = k, k+1, ..., k_1-1$, и $J_{k_1} = I_k^1$. Тогда

(2.21)
$$|J_m| = \frac{1}{2} |I_m^1| = 2^{-m-1} = |I_m^*|.$$

Нетрудно заметить, что сели $x \in J_m$, то из (2.20) следует

$$\Sigma_2(x) \leq C \cdot \sum_{m=k}^{k_1} \varphi_m(x) \leq C \cdot (B_k + \cdots + B_m + B_{m+1} \cdot q + \cdots + B_{k_1} \cdot q^{k_1 - m}).$$

Поэтому

(2.22)
$$\int_{I_{k}^{1}} (\Sigma_{2}(x))^{p} dx = \sum_{j=1}^{k} \int_{\mathbb{R}^{n}} (\Sigma_{2}(x))^{p} dx \le$$

$$C_p \cdot \sum_{m=k}^{k_1} |J_m| (\sum_{\nu=k}^m B_{\nu})^p + C_p \cdot \sum_{m=k}^{k_1} |J_m| (\sum_{\nu=m+1}^{k_1} B_{\nu} \cdot \eta^{-m})^p =: C_p(\Gamma_1 + \Gamma_2).$$

Пусть $q_1 \in (0,1)$, такое что $q_1^{-p} < 2$. Тогда для Σ_1 получим

$$\Gamma_1 = C_p \sum_{m=k}^{k_1} |J_m| \left(\sum_{\nu=k}^m B_\nu q_1^{\nu-k} q_1^{k-\nu} \right)^p \le C_p \sum_{m=k}^{k_1} |J_m| \sum_{\nu=k}^m B_\nu^p q_1^{\nu(\nu-k)} \left(\sum_{\nu=k}^m q_1^{\nu(\nu-k)} \right)^{p-1}.$$

Учитывая, что в последней сумме $\nu \geq k$ и $q_1 < 1$, имеем

$$\left(\sum_{p=k}^m q_1^{\frac{p(k-p)}{p-1}}\right)^{p-1} \leq C_p \cdot q_1^{p(k-m)}.$$

Поэтому, для Γ_1 имеем

(2.23)
$$\Gamma_1 \le C_p \sum_{m=k}^{k_1} |J_m| \sum_{\nu=k}^m B_{\nu}^p \cdot q_1^{p(\nu-m)} = C_p \sum_{\nu=k}^{k_1} B_{\nu}^p \sum_{m=\nu}^{k_1} |J_m| \cdot q_1^{p(\nu-m)}.$$

Так как $|J_{m+1}|=\frac{1}{2}|J_m|$ и $q_1^{-p}<2$, то $|J_{m+1}|\cdot q_1^{p(n-m-1)}<\gamma\cdot |J_m|\cdot q_1^{p(n-m)}$, для некоторого $\gamma<1$. Поэтому из (2.23) и (2.18) имеем

(2.24)
$$\Gamma_1 \le C_p \sum_{\nu=k}^{k_1} B^p_{\nu} |J_{\nu}| < C_p, \text{ Kormal } p > 1.$$

Оценим Γ_2 , когда p>1. Обозначим $q_1=0.9$ и напомним, что q=2/7. Тогда, учитывая что $q\cdot q_1^{-1}<1$, из (2.22) с применением неравенства Гелдера получим

(2.25)
$$\Gamma_{2} = C_{p} \sum_{m=k}^{k_{1}} |J_{m}| \left(\sum_{\nu=m+1}^{k_{1}} B_{\nu} \cdot (q_{1}^{-1} \cdot q)^{\nu-m} \cdot q_{1}^{\nu-m} \right)^{p} \leq C_{p} \sum_{m=k}^{k_{1}} |J_{m}| \left(\sum_{\nu=m+1}^{k_{1}} B_{\nu}^{p} \cdot (q_{1}^{-1} \cdot q)^{p(\nu-m)} \right) \cdot \left(\sum_{\nu=m+1}^{k_{1}} q_{1}^{(\nu-m)}^{p-1} \right)^{p-1} \leq C_{p} \sum_{m=k}^{k_{1}} |J_{m}| \sum_{\nu=m+1}^{k_{1}} B_{\nu}^{p} \cdot (q_{1}^{-1} \cdot q)^{p(\nu-m)} = C_{p} \sum_{m=k}^{k_{1}} B_{\nu}^{p} \sum_{\nu=1}^{p-1} |J_{m}| (q_{1}^{-1} \cdot q)^{p(\nu-m)}$$

Из (2.25) и (2.21) следует

$$\Gamma_2 \leq C_p \sum_{\nu=k+1}^{k_1} B_{\nu}^p \cdot |J_{\nu}| \sum_{m=k}^{\nu-1} 2^{\nu-m} (q_1^{-1} \cdot q)^{p(\nu-m)}$$

Учитывая, что $2\cdot (q_1^{-1}\cdot q)^p<2\cdot \frac{10}{9}\cdot \frac{2}{7}<1$. из (2.25), (2.21) и (2.18) получим

(2.26)
$$\Gamma_2 \le C_p \sum_{\nu=k}^{k_1} B_{\nu}^p |I_{\nu}^*| \le C_p.$$

Из (2.26). (2.24) и (2.22) получаем

$$\int_{\mathbb{R}} \left(\Sigma_2(x) \right)^p \! dx \leq C_p, \quad \text{когда} \quad p > 1.$$

Очевидно, из (2.20) имеем

$$\Sigma_2(x) \leq C \cdot \sum_{m=k}^{k_1} \varphi_m(x) \leq C \cdot q^{j-1} \sum_{m=k}^{k_1} \varphi_m\left(x+rac{j}{2^{k+1}}
ight),$$
 когда $x \in I_k^j$.

Следовательно, получаем

$$\int_{P_b^1} (\Sigma_2(x))^p dx \leq C_p \cdot q^{j-1} \int_{P_b^1} (\Sigma_2(x))^p dx.$$

Отсюда и из (2.27) получим (2.13) для i=2.

В случае i=1 неравенство (2.13) доказывается чуть проще. В этом случае нет необходимости соотношений типа (2.14), (2.15), поскольку $\sigma_m(x)$ одна функция о которой известно, что (см. (2.5))

$$\max_{x \in I_{-}^{2}} \sigma_{m}(x) \leq \left(\frac{2}{7}\right)^{j-1} \cdot \sigma_{m}\left(\frac{w}{2^{k}}\right) \leq C \cdot \left(\frac{2}{7}\right)^{j-1} \cdot A_{m},$$

где отрезки I_m^J те же, что в соотношении (2.20). Второе неравенство следует из того, что $\infty > 0$ (в противном случае нечего доказывать) и поэтому (см. Предложение 2.2 и (2.16), (2.17))

$$\sigma_m\left(\frac{x}{2^k}\right) \le \max_{x \in I_m} \sigma_m(x) = \max_{x \in I_m^x} \sigma_m(x) \le C \cdot A_m$$

Отсюда, рассуждениями, аналогичными рассуждениям при доказательстве (2.27), получим

$$\int_0^{\frac{2n}{2^k}} (\Gamma_1(x))^p dx \le C_p, \text{ для } p > 0.$$

Тем самым доказано неравенство (2.11). Аналогично доказывается, что

$$\int_{\frac{w+1}{2k}}^{1} (S^{\mathbf{v}}(x))^p dx < C_p.$$

которое вместе с (2.11) доказывают неравенство (2.3). Перавенство (2.4) доказывается аналогично. Лемма 2.1 доказана.

Доказательство теоремы 2.2. Сначала убедимся в справедливости неравенства (2.1). Поскольку $\left[\frac{w}{2^k}, \dots\right] \neq [0,1]$, то либо x > 0 либо $x < 2^k$. Рассмотрим случай, когда x > 0. Для x > 0 для x >

(2.28)
$$a_{2^m+l}f_{2^m+l}\left(\frac{ie}{2^k}\right) = \frac{1}{l}, \text{ korga } l = 1, 2, ..., x \cdot 2^{r_1 - k}$$

Убедимся, что для любого C>0 при достаточно больтом m ряд $\sum_{n=0}^{m-2}a_{2^m+1}f_{2^m+1}(x)$ удовлетворяет (2.1). Во первых

$$(2.29) \left\| \sup_{N} \left| \sum_{n \le N} a_n f_n(x) \right| \right\|_{L_p(I)} > \int_{\frac{\infty}{2^k}}^{\frac{\infty}{2^k} + \frac{1}{2^m}} \left| \sum_{l=1}^{\infty \cdot 2^{m-k}} a_{2^m + l} f_{2^m + l}(x) \right|^p dx > C_0 \frac{m - k}{2^m}.$$

С другой стороны (см. (2.28) и (2.6))

$$(2.30) \qquad \left\| \left\{ \sum_{n} a_{n}^{2} f_{n}^{2}(x) \right\}^{\frac{1}{2}} \right\|_{L_{p}(I)} \leq \int_{\frac{\pi}{2^{k}}}^{\frac{\kappa}{2^{k}} + \frac{1}{2^{m}}} \left(\sum_{n} a_{2^{m}+l}^{2} f_{2^{m}+l}^{2}(x) \right)^{\frac{p}{2}} dx \sum_{\eta=0}^{\infty} q^{\eta p} \leq C_{p} \cdot 2^{-m} \left(\sum_{n} a_{2^{m}+l}^{2} f_{2^{m}+l}^{2} \left(\frac{\pi}{\alpha^{k}} \right) \right)^{\frac{p}{2}} \leq C_{p} \cdot 2^{-m}.$$

При достаточно большом m из (2.29) и (2.30) следует (2.1).

Случай $x < 2^k$ доказывается аналогично. Только в сумме $\sum_n a_n f_n(x)$ будут участвовать n с условиями [n] = m, $\{n\}$ находится правее I.

Докажем соотношение (2.2). Опить рассмотрим только случай x>0. Для m>k и $l\leq x\cdot 2^{m-k}$ выберем a_{2^m+l} так. чтобы

(2.31)
$$a_{2^m+l}f_{2^m+l}\left(\frac{w}{2^k}\right) = (-1)^l$$
, когда $l = 1, 2, ..., x \cdot 2^{m-k}$.

Убедимся, что для любого C>0 при достаточно большом m ряд $\sum_n a_n f_n(x):=\sum_{l=1}^{n\cdot 2^{m-k}}a_{2^m+l}f_{2^m+l}(x)$ удовлетворяет (2.2).

Из Предложения 2.4 и (2.31) следует, что

$$\sup_{n \le N} \left| \sum_{n \le N} a_n f_n(x) \right| = |a_{2^m+1} f_{2^m+1}(x)|, \quad \text{когда} \quad x \ge \frac{x}{2^k}.$$

Поэтому

(2.32)
$$\left\| \sup_{N} \left| \sum_{n \le N} a_n f_n(x) \right| \right\|_{L_p(I)}^p < \int_{I} |a_{2^m+1} f_{2^m+1}(x)|^p dx < C_p \cdot 2^{-m}.$$

С другой стороны

(2.33)

$$\int_{I} \left(\sum a_{n}^{2} f_{n}^{2}(x) \right)^{-} dx > \int_{1}^{\frac{w}{2^{k}} + \frac{1}{2^{m}}} \left(\sum_{l=1}^{m2^{k}} a_{2^{m}+l}^{2} f_{2^{m}+1}^{*}(x) \right)^{-} dx > C_{p}(\varpi 2^{m-k})^{\frac{p}{2}} 2^{-m}$$

Из (2.32) и (2.33), при достаточно большом m, следует (2.2). Теорема 2.2 доказана.

Доказательство теоремы 2.3. Допустим

$$I = \left\lfloor \frac{\infty}{2^k}, \frac{m+1}{2^k} \right\rfloor \quad \text{if } \|S^*(\cdot)\|_{L_p(I)} = 1, \quad \text{fige} \quad S^*(x) = \sup_N \left| \sum_{n \leq N} a_n \chi_n(x) \right|$$

Пусть $n_1 < n_2 < ... < n_k$, те номера, для которых

$$\Delta_{n_i} \cap I \neq \emptyset$$
, if $\Delta_{n_i} \not\subset I$, the $\Delta_n = \operatorname{supp} \chi_n$.

Отметим, что таких n_i ровно k+1 штук. Если учесть, что функции $a_{n_i}\chi_{n_i}(x)$ принимают постоянные значения на I, то получим

$$\max_{x\in I}|a_{n_i}\chi_{n_i}|\leq 2\min_{x\in I}S^*(x)\leq 2S^*(x)\quad \text{if}\quad \max_{I}\left|\sum_{x=1}^{I}a_{n_i}\chi_{n_i}(x)\right|\leq S^*(x),\quad \text{for}\quad x\in I.$$

Следовательно, получаем

(2.34)
$$\sum_{n=1}^{\infty} a_n^2 f_n^2(x) \le 4(k+1) \cdot (S^*(x))^2 + \sum_{n \Delta_n \subset I} a_n^2 f_n^2(x) \quad x \in I$$

H

(2.35)
$$S_1^*(x) := \sup_{N} \left| \sum_{\Delta_n \in I: n \le N} a_n \chi_n(x) \right| \le 2 \cdot S^*(x), \quad x \in I.$$

Учитывая, что

$$\left\| \sum_{n:\Delta_n\subset I} a_n^2 \chi_n^2(\cdot) \right\|_{L_p(I)} \sim_{\mathfrak{p}} \|S_1''(\cdot)\|_{L_p(I)},$$

из (2.34), (2.35) получим $\int_{\mathbb{R}} \left(\sum_{n=1}^{\infty} a^2 \chi_n^2(x) \right)^{\frac{1}{2}} dx \leq C_{p,k} \int_{\mathbb{R}} (S^*(x))^p dx$.

Аналогично доказывается неравенство $\int_I (S^*(x))^p dx \le C_{p,k} \int_I \left(\sum_{n=1}^\infty a_n^2 \chi_n^2(x)\right)^{\frac{p}{2}} dx.$ Тем самым доказали следующее соотношение

$$\int_I (S^*(x))^p dx \sim_{p,k} \int_I \left(\sum_{n=1}^\infty a_n^2 \chi_n^2(x) \right)^{\frac{1}{p}} dx,$$

причем постоянные эквивалентности зависят не от самой I_* а от ранга I_* Теорема доказана.

В теоремах 2.1 и 2.3, вообще говоря, мажоранту ряда нельзя заменить суммой ряда. Действительно, известно, что в пространстве L[0,1] не существует безусловных базисов (см. [11]). Отсюда следует, что существуют функции $\phi,\psi\in L_1$, такие, что

$$\left(\sum_{n=0}^{\infty}a_{n}^{2}f_{n}^{2}(\cdot)
ight)^{\frac{1}{2}}$$
 = L_{1} где $a_{n}=\int_{0}^{1}\phi(x)f_{n}(x)dx$.

И

$$\left(\sum_{n=0}^{\infty}b_{n}^{2}\chi_{n}(\cdot)\right)$$
 $otin L_{1}, ext{ fig. } b_{n}=\int_{0}^{1}\psi(x)\chi_{n}(x)dx.$

Следовательно

$$\| \sum_n a_n \varphi_n \|_1 \not \sim \| (\sum_n a_n^2 \varphi_n^2)^{\frac{1}{2}} \|_1.$$

где $\{\varphi_n\}$ -система Хаара пли Франклина.

В случае p>1, из безусловной базисности системы Хаяра в пространстве L_p и конструкции функций Хаара, для любого двоичного интервала I имеем

(2.36)
$$\left\| \sum_{\Delta_n \subset I} a_n \chi_n \right\|_{L_p(I)} \sim_p \left\| \left(\sum_{\Delta_n \subset I} a_n^2 \chi_n^2 \right)^{\frac{1}{2}} \right\|_{L_p(I)}$$

Интересно было бы выяснить, имеет ли аналог соотношения (2.36) для системы Франклина. Для установления такой эквивалентности необходимо (и достаточно) установить аналог соотношений (2.3), (2.4) для суммы $\sum_{\{n\}\subset I} a_n f_n(x)$.

Abstract. In this paper we prove that the majorant of partial sums and the Paley function of Franklin series have equivalent norms in the space $L_p(I)$, p > 0, provided that the "peak" intervals of Franklin functions with non-vanishing coefficients lie in I. Examples of series emphasizing that this condition is essential are also given.

Список литературы

- [1] Ph. Franklin, "A set of continuos orthogonal functions", Math. Ann. 100, 522 528 (1928).
- [2] Z. Ciesielski, "Properties of the orthonormal Franklin system", Studia Math. 23, 141 157 (1963).
- [3] Z. Ciesielski, "Properties of the orthonormal Franklin system IΓ". Studia Math. 27, 289 323 (1966).
- [4] S. V. Bochkarev, "Some inequalities for the Franklin series", Anal. Math. 1, 249 257 (1975).
- [5] Б. С. Кашин, А. А. Саакян, Ортоговальные Ряды, Москва, АФЦ (1999).
- [6] P. Sjolin, "Convergence almost everywhere of spline expansions in Hardy spaces". In: Topics in modern harmonic analysis. (Turin/Milan, 1982). Ist. Naz. Alta Mat. Francesco Severi, Rome, 645 - 651 (1983).
- [7] P. Sjolin, J. O. Stromberg, "Basis properties of Hardy spaces", Ark. Mat. 21, 111 125 (1983).
- [8] Г. Г. Геворкян. "Некоторые теоремы о безусловной скодимости и мажоранте рядов Франжина и их применение к пространствам Re(H^p)." Тр. MMAH CCCP, 190, 49 – 74 (1989). Translation: G. G. Gevorkian, Some theorems on unconditional convergence and a majorant of Franklin series and their application to the spaces Re(H^p). Proceedings of the Steklov Institute of Mathematics, 49 – 76 (1992).
- [9] Г. Г. Геворкян, "О рядах по системе Франклина". Anal. Math., 16, 87 114 (1990).
- [10] Г. Г. Геворкян, "Неограниченность оператора сдвига по системе Франклина в пространстве L_1 ", Мат. Заметки, 38, по. 4, 523 533 (1985).
- [11] A. Pelczynski, "Projections in certain Banach spaces". Studia Math. 19, 209 228 (1960).

Поступила 18 мая 2016

Известия НАН Арменци, Математика, том 52, п. 1, 2017, стр. 59-67

ON L'-INTEGRABILITY OF A SPECIAL DOUBLE SINE SERIES FORMED BY ITS BLOCKS

XII. Z. KRASNIQI

University of Prishtina, Prishtina, Kosovo E-mail: xhevat.krasniqi@uni-pr.edu

Abstract. In this paper we deal with a special double sine trigonometric series formed by its blocks. This type of trigonometric series is of particular interest since its blocks always are bounded, that is, under some additional assumptions the sumfunction of such series always exists. We give some conditions under which such sumfunction is integrable of power $p \in \{2, 3, \dots\}$, as well as is integrable with some natural weight.

MSC2010 numbers: 42A16, 42A20, 28A25,

Keywords: Sine series; function of bounded variation; series by their blocks.

1. Introduction

Let $\Lambda_1 = \{n_1\}$ and $\Lambda_2 = \{r_2\}$ be two strictly increasing sequences of natural numbers $1 = n_1 < n_2 < n_3 < \cdots$ and $1 = r_1 < r_2 < r_3 < \cdots$ satisfying the conditions:

$$\sum_{i=1}^{\infty} \frac{1}{n_i} < +\infty \quad \text{and} \quad \sum_{i=1}^{\infty} \frac{1}{r_i} < +\infty.$$

Considering the special double sine series

$$\sum_{k=1}^{\infty} \sum_{\ell=1}^{\infty} \frac{\sin kx \sin \ell y}{k\ell}$$

we form the following series

(1.1)
$$\sum_{r=1}^{\infty} \sum_{j=1}^{\infty} \left| \sum_{k=n_j}^{n_{i+1}-1} \sum_{\ell=r_j}^{r_{j+1}-1} \frac{\sin k x \sin \ell y}{k \ell} \right|.$$

According to the well-known estimate

(1.2)
$$\left| \sum_{k=0}^{V} \frac{\sin kx}{k} \right| \le \frac{\pi}{vx}, \quad v \le V \le \infty, \quad 0 < x \le \pi.$$

the series (1.1) converges for all (x, y) and its sum $G_{\Lambda_1, \Lambda_2}(x, y)$ is a continuous function on $(0, \pi] \times (0, \pi]$. This fact is of particular interest and therefore this is the main reason why we have formed the series (1.1).

In the one-dimensional case such series has been considered by Telyakovskii [1] and Trigub [3]. In particular, Telyakovskii [2] has considered the question: when the sum-function $g_{\Lambda_1}(x)$ of the series

$$\sum_{k=1}^{\infty} \left| \sum_{k=n_k}^{n_{k+1}-1} \frac{\sin kx}{k} \right|$$

belongs to the spaces $L^p[0,\pi]$ for $p=2,3,\ldots$?

Specifically, in [2] was proved the following theorem.

Theorem 1.1. For any natural p = 2, 3, ... the function $g_{\Lambda_1}(x)$ belongs to the space $A^{(p)}[0, \pi]$ if the series $\sum_{i=1}^{n} \frac{1}{m}$ is convergent, where $m_i = \min(n_i, n_{i+1} - n_i + 1)$.

In the same paper was considered the problem of integrability of the function $g_{\Lambda_1}(x)$ with weight $x^{-\gamma}$ under natural condition $0 < \gamma < 1$. Among others, the following result was proved in [2].

Theorem 1.2. If for $\gamma \in (0,1)$ the series

$$\sum_{i=1}^{\infty} \frac{1}{n_i} m_i^{\gamma}$$

is convergent, then the integral $\int_0^{\infty} \frac{1}{x^{\gamma}} g_{\Lambda_1}(x) dx$ converges.

Note that questions pertaining to trigonometric series formed by their blocks were considered in [4] - [6], and still receive considerable attention. The main aim of this paper is to extend the above results to two-dimensional case. In order to do this we will use the technique developed in [2], the estimate (1.2) and the following inequality (see [2] page 818):

(1.3)
$$u_i(x) := \left| \sum_{k=n_i}^{n_{i+1}-1} \frac{\sin kx}{k} \right| \le \frac{A}{n_i} \min \left(\frac{1}{x}, m_i \right), \quad 0 < x \le \pi.$$

2. THE MAIN RESULTS

In this section we state and prove the main results of the paper. We first prove the following result.

Theorem 2.1. For any natural p = 2, 3, ... the function G_{Λ_1, Λ_2} belongs to the space $L^p([0, \pi] \times [0, \pi])$ if the series

$$\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{(m_i s_j)^{1-\frac{1}{p}}}{n_i r_j}$$

is convergent, where $m_i = \min(n_i, n_{i+1} - n_i + 1)$ and $s_j = \min(r_j, r_{j+1} - r_j + 1)$.

Proof. For arbitrary natural numbers M and N we have

$$\int_{0}^{\pi} \int_{0}^{\pi} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} u_{i}(x) u_{j}(y) \right)^{p} dx dy$$

$$= \int_{0}^{\pi} \int_{0}^{\pi} \sum_{i=1}^{M} u_{i_{1}}(x) \cdots \sum_{i_{p}=1}^{M} u_{i_{p}}(x) \sum_{j_{1}=1}^{N} u_{j_{1}}(y) \cdots \sum_{j_{p}=1}^{N} u_{j_{p}}(y) dx dy$$

$$= \sum_{i_{1}=1}^{M} \cdots \sum_{i_{p}=1}^{M} \sum_{j_{1}=1}^{N} \cdots \sum_{j_{p}=1}^{N} \int_{0}^{\pi} \int_{0} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy.$$
(2.1)

Next, we split the square $[0,\pi] \times [0,\pi]$ into the rectangles $[0,\alpha] \times [0,\beta]$, $[0,\alpha] \times [\pi,\beta]$, $[\alpha,\pi] \times [0,\beta]$ and $[\alpha,\pi] \times [\beta,\pi]$, where α and β will be determined later in an appropriate way. Using the estimates (1.3) we can write

$$\int_{0}^{\alpha} \int_{0}^{\beta} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$(2.2) \leq A^{2p} \int_{0}^{\alpha} \int_{0}^{\beta} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{s_{p}} dx dy = A^{2p} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{s_{j_{p}}}{n_{i_{p}}} r_{j_{1}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \alpha \beta,$$

$$(2.3) \int_{0}^{\alpha} \int_{0}^{\pi} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{p}}(y) dx dy$$

$$\leq A^{2} \int_{0}^{\alpha} \int_{0}^{\pi} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{1}{r_{j_{1}} \cdots r_{j_{p}}} \frac{dx dy}{y} \leq \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{A^{2p}}{r_{j_{1}} \cdots r_{j_{p}}} \frac{\alpha \beta^{1-p}}{p-1}.$$

$$(2.4) \qquad \int_{0}^{\pi} \int_{0}^{\beta} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$\leq A^{2p} \int_{0}^{\beta} \int_{0}^{\beta} \frac{1}{n_{i_{1}} \cdots n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{dx dy}{x^{p}} \leq \frac{A^{2p}}{n_{i_{1}} \cdots n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{\alpha^{1-p} \beta}{p-1},$$

(2.5)
$$\int_{0}^{\infty} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$\leq A^{2p} \int_{0}^{\infty} \int_{0}^{\infty} \frac{1}{n_{i_{1}} \cdots n_{i_{p}}} \frac{1}{r_{j_{1}} \cdots r_{j_{p}}} \frac{dx dy}{(xy)^{p}} < \frac{A^{2p}}{n_{i_{1}} \cdots n_{i_{p}} r_{j_{1}} \cdots r_{j_{p}}} \frac{(\alpha \beta)^{1-p}}{(p-1)^{2}}$$

$$61$$

Inserting the estimates (2.2)-(2.5) into (2.1), we obtain

$$\int_{0}^{\pi} \int_{0}^{\pi} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} u_{i}(x) u_{j}(y) \right)^{p} dx dy < A^{2p} \sum_{i_{1}=1}^{M} \cdots \sum_{i_{n}=1}^{M} \sum_{i_{1}=1}^{N} \sum_{i_{1}=1}^{N} \sum_{i_{2}=1}^{N} \left(\frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{s_{j_{1}}}{r_{j_{2}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \cdots \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \cdots \frac{s_{j_{p}}}{r_{p}}} \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \cdots \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \frac{s_{j_{p}}}{r_{p}} \cdots \frac{s_{j_{p}}}{r_{p}} \frac{s$$

Whence, choosing in (2.6) $\alpha = (m_{i_1} \cdots m_{i_p})^{-\frac{1}{p}}$ and $\beta = (s_{j_1} \cdots s_j)^{-\frac{1}{p}}$ we find that

$$\int_{0}^{\pi} \int_{0}^{\cdot} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} n_{i}(x) n_{j}(y) \right)^{p} dx dy < 4A^{2p} \sum_{i_{1}=1}^{M} \cdots \sum_{i_{p}=1}^{M} \sum_{j_{1}=1}^{N} \cdots \sum_{i_{p}=1}^{M} \sum_{j_{1}=1}^{N} \sum_{j_{1}=1}^{N} \frac{(m_{i_{1}} \cdots m_{i_{p}} s_{j_{1}} \cdots s_{j_{p}})^{1-\frac{1}{p}}}{n_{i_{1}} \cdots n_{i_{p}} r_{j_{1}} \cdots r_{j_{p}}} < 4A^{2p} \left(\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{(m_{i} s_{j})^{1-\frac{1}{p}}}{n_{i} r_{j}} \right)^{p}.$$

Consequently, since the last series converges by assumption, the integrals

$$\int_0^\pi \int_0^\pi \left(\sum_{i=1}^M \sum_{j=1}^N u_i(x) u_j(y) \right)^p dx dy$$

are bounded by a quantity that is independent of M,N. Therefore, based on the double version of the Levi's theorem, we conclude that the function G_{Λ_1,Λ_2} belongs to the space $L^p([0,\pi]\times[0,\pi])$.

The next result gives an answer to the following question: under what conditions the function G_{Λ_1,Λ_2} belongs to the space $L^p([0,\pi]\times[0,\pi])$ with weight $x^{-\gamma_1}y^{-\gamma_2}$, $\gamma_1,\gamma_2\in(0,1)$?

Theorem 2.2. If for $\gamma_1, \gamma_2 \in (0,1)$, the series $\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{1}{n} m_i^{\gamma_j} s_j^{\gamma_2}$ is convergent, then the following integral converges

$$\int_0^\pi\!\int_0^\pi \frac{G_{\Lambda_1,\Lambda_2}(x,y)}{x^{\gamma_1}y^{\gamma_2}}dxdy$$

Proof. Based on the uniform convergence of the series (1.1) we have

$$\int_0^\pi\!\int_0^\pi \frac{G_{\Lambda_1,\Lambda_2}(x,y)}{x}dxdy = \sum_{i=1}^\infty \sum_{j=1}^\infty \int_0^\pi\!\int_0^\pi \frac{u_i(x)u_j(y)}{x}dxdy.$$

Splitting the square $[0, \pi] \times [0, \pi]$ into the rectangles $[0, \alpha_i] \times [0, \beta_j]$ $[0, \alpha_i] \times [\pi, \beta_j]$, $[\alpha_i, \pi] \times [0, \beta_j]$ and $[\alpha_i, \pi] \times [\beta_j, \pi]$, where α_i and β_j are determined by

(2.7)
$$\alpha_{\epsilon} = \frac{1}{m} \quad \text{and} \quad \beta_{j} = \frac{1}{s_{j}}$$

we find that

$$\int_{0}^{\alpha_{i}} \int_{0}^{\beta_{j}} \frac{u_{i}(x)u_{j}(y)}{x^{\gamma_{i}}y^{\gamma_{2}}} dxdy \leq A^{2} \int_{0}^{\alpha_{i}} \int_{0}^{\beta_{j}} \frac{1}{x^{\gamma_{i}}y^{\gamma_{2}}} \frac{m_{i}s_{j}}{n_{i}r_{j}} dxdy$$

$$= \frac{1}{(1 - \gamma_{1})(1 - \gamma_{2})} \frac{m_{i}s_{j}}{n_{i}r_{j}} dxdy$$

$$\int_{\alpha_{i}} \int_{0}^{\beta_{j}} \frac{u_{i}(x)u_{j}(y)}{x^{\gamma_{i}}y^{\gamma_{2}}} dxdy \leq A^{2} \int_{\alpha_{i}}^{\pi} \int_{0}^{\beta_{j}} \frac{1}{x^{\gamma_{j}}y^{\gamma_{2}}} \frac{s_{j}}{n_{i}xr_{j}} dxdy$$

$$= \frac{1}{\gamma_{1}(1 - \gamma_{2})} \frac{1}{n_{i}r_{j}} \frac{s_{j}}{n_{i}xr_{j}} dxdy$$

$$= \frac{1}{\gamma_{1}(1 - \gamma_{2})} \frac{1}{n_{i}r_{j}} \frac{s_{j}}{n_{i}r_{j}y} \frac{1}{n_{i}r_{j}y}$$

$$= \frac{1}{(1 - \gamma_{1})\gamma_{2}} \frac{m_{i}}{n_{i}r_{j}} \frac{1}{n_{i}r_{j}y}$$

$$\int_{\alpha_{i}}^{\pi}\!\!\int_{\beta_{j}}^{\pi}\frac{u_{i}(x)u_{j}(y)}{x^{\gamma_{1}}y^{\gamma_{2}}}dxdy \leq A^{2}\int_{\alpha_{i}}^{\pi}\!\!\int_{\beta_{j}}^{\pi}\frac{1}{x^{\gamma_{1}}y^{\gamma_{2}}}\frac{1}{n_{i}r_{j}xy}dxdy \leq \frac{A^{2}}{\gamma_{1}\gamma_{2}}\frac{1}{n_{i}r_{j}}dxdy \leq \frac{A^{2}}{\gamma_{1}\gamma_{2}}\frac{1}{n_{i}r_{j}}dxd$$

Finally, using (2.7) and the latest estimates, we obtain

$$\int_{0}^{\infty} \int_{0}^{\infty} \frac{G_{\Lambda_{1}\Lambda_{2}}(x,y)}{x^{\gamma_{1}}y^{\gamma_{2}}} dxdy < \sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \left(\frac{A^{2}}{(1-\gamma_{1})(1-\gamma_{2})} \frac{m_{i}s_{j}}{n_{i}r_{j}} n^{1-\gamma_{1}} \beta^{1-\gamma_{2}} + \frac{A^{2}}{(1-\gamma_{1})\gamma_{2}} \frac{m_{i}s_{j}}{n_{i}r_{j}} \alpha^{1-\gamma_{1}} \beta^{-\gamma_{2}} + \frac{A^{2}}{\gamma_{1}\gamma_{2}} \frac{m_{i}s_{j}}{n_{i}r_{j}} \alpha^{1-\gamma_{1}} \beta^{-\gamma_{2}} \right) = C \sum_{j=1}^{\infty} \sum_{j=1}^{\infty} \frac{1}{n_{j}} m^{\gamma_{1}} s_{j}^{\gamma_{2}} < +\infty.$$

where
$$C = A^2 \cdot \max \left\{ \frac{1}{1 - m(1 - m)}, \frac{1}{1 - m(1 - m)}, \frac{1}{1 - m}, \frac{1}{1$$

The next statement supplements Theorem 2.1, and gives conditions under which the integral

$$\int_0^\pi \! \int_0^\pi \frac{G_{\Lambda_1,\Lambda_2}(x,y)}{x^{\gamma_1}y^{\gamma_2}} dx dy$$

is convergent for $\gamma_1, \gamma_2 \in (0, 1)$ and p = 2, 3, ...

Theorem 2.3. If $p = 2, 3, \ldots$ and $\gamma_1, \gamma_2 \in (1 - p, 1)$, then the integral

$$\int_0^\pi \int_0^\pi \frac{G^p_{\Lambda_1,\Lambda_2}(x,y)}{x^{\gamma_1}y^{\gamma_2}} dxdy$$

is convergent provided that the series

$$\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{1}{n_i r_j} n_{ij}^{1-\frac{1}{n}(1-\gamma_1)} n_{-\frac{1}{n}(1-\gamma_2)}$$

is convergent.

Proof. Using a similar technique as in the proof Theorem 2.1 we have

$$\int_{0}^{x} \int_{0}^{x} \frac{1}{x^{\gamma_{1}}y^{\gamma_{2}}} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} u_{i}(x)u_{j}(y) \right)^{n} dx dy$$

$$(2.8) = \sum_{j_{1}=1}^{M} \cdots \sum_{j_{n}=1}^{M} \sum_{j_{1}=1}^{N} \cdots \sum_{j_{n}=1}^{N} \int_{0}^{x} \int_{0}^{x} \frac{1}{x^{\gamma_{1}}y^{\gamma_{2}}} u_{i_{1}}(x) \cdots u_{i_{p}}(x)u_{j_{1}}(y) \cdots u_{j_{p}}(y) dxdy,$$

for all $p = 2, 3, \ldots$ and natural numbers M, N

Again we split the square $[0, \pi] \times [0, \pi]$ into the rectangles $[0, \alpha] \times [0, \beta]$, $[0, \alpha] \times [\pi, \beta]$, $[\alpha, \pi] \times [0, \beta]$ and $[\alpha, \pi] \times [\beta, \pi]$, where α and β are determined as in Theorem 2.1.

Using the estimates (1.3) and taking into account that $\gamma_1 \chi \gamma_2 \in (1 + p, 1)$, we can write

(2.9)
$$\int_{0}^{\infty} \int_{0}^{1} \frac{1}{x^{j_{1}} u_{j_{1}}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) axdy$$

$$\leq A^{2p} \int_{0}^{\alpha} \int_{0}^{\beta} \frac{1}{x^{j_{1}} u_{j_{1}}} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} r_{j_{1}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} dxdy$$

$$= A^{2p} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{s_{j_{p}}}{1} \frac{\alpha^{1-\gamma_{1}} \beta^{1-\gamma_{2}}}{(1-\gamma_{1})(1-\gamma_{2})}$$

$$\int_{0}^{\pi} \int_{\beta}^{\pi} \frac{1}{\prod_{i=1}^{n} (x) \cdots u_{i_{\nu}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dxdy} \\
\leq A^{2p} \int_{0}^{\alpha} \int_{\beta}^{\pi} \frac{m_{i_{1}}}{n_{i_{1}}} \frac{m_{i_{p}}}{n_{i_{p}}} \frac{1}{r_{j_{1}} \cdots r_{j_{p}}} \frac{dxdy}{r^{j_{1}} \cdots r_{j_{p}}} \\
< \frac{n_{i_{1}}}{n_{i_{1}}} \frac{m_{i_{p}}}{n_{i_{p}}} \frac{A^{2p}}{\prod_{i=1}^{n} (1 - \gamma_{1})(\gamma_{2} + p - 1)} \\$$
(2.10)

$$\int_{\alpha} \int_{0}^{\pi} \frac{1}{x_{1} + x_{2}} u_{j_{1}}(x) \dots u_{j_{p}}(x) u_{j_{1}}(y) \dots u_{j_{p}}(y) dx dy \\
\leq A^{2p} \int_{0}^{\pi} \int_{0}^{\beta} \frac{1}{n_{i_{1}} \dots n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \dots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{dx dy}{x_{j_{1} + py_{j_{1}}}} \\
\leq \frac{A^{2p}}{n_{i_{1}} \dots n_{i_{p}}} \frac{1}{r_{j_{p}}} \frac{\alpha^{1 - \gamma_{1} - p} \beta^{1 - \gamma_{2}}}{(\gamma_{1} + p - 1)(1 - \gamma_{2})}, \\
64$$

ON LP-INTEGRABILITY OF A SPECIAL DOUBLE SINE

and

$$\int_{\alpha} \int_{\beta} \frac{1}{x^{\gamma_{1}} \eta^{\gamma_{2}}} u_{i_{1}}(x) = u_{i_{1}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dxdy \\
\leq A^{2p} \int_{\alpha}^{\pi} \int_{\beta}^{\pi} \frac{1}{n_{i_{1}} \cdots n_{i_{p}}} \frac{1}{r_{j_{1}} \cdots r_{j_{p}}} \frac{dxdy}{x^{\gamma_{1} + p} \eta^{\gamma_{2} + p}} \\
= \frac{A^{2p}}{n_{i_{1}} \cdots n_{i_{p}} r_{j_{1}} \cdots r_{j_{p}}} \frac{\alpha^{1 - \gamma_{1} - p} \beta^{1 - \gamma_{2} - p}}{(\gamma_{1} + p - 1)(\gamma_{2} + p - 1)}.$$

The above estimates along with

$$\alpha = \frac{1}{(m_{i_1} \cdots m_{i_p})^{\frac{1}{p}}}$$
 and $\beta = \frac{1}{(s_{j_1} \cdots s_{j_p})^{\frac{1}{p}}}$

iniply

$$\begin{split} & \int_{0}^{\pi} \int_{0}^{\pi} \frac{1}{x^{\gamma_{1}} y^{\gamma_{2}}} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} u_{i}(x) u_{j}(y) \right)^{p} dx dy \\ & < A(p, \gamma_{1}, \gamma_{2}) \sum_{i_{1}=1}^{M} \cdots \sum_{i_{p}=1}^{M} \sum_{j_{1}=1}^{N} \cdots \sum_{j_{p}=1}^{N} \frac{(m_{i_{1}} \cdots m_{i_{p}})^{1-\frac{1}{p}(1-\gamma_{1})} (s_{j_{1}} \cdots s_{j_{p}})^{1-\frac{1}{p}(1-\gamma_{1})}}{n_{i_{1}} \cdots n_{i_{p}} r_{j_{1}} \cdots r_{j_{p}}}. \end{split}$$

where $A(p, \gamma_1, \gamma_2)$ is a constant that depends only on p, γ_1 , and γ_2 .

Hence,

$$\begin{split} \int_{0}^{\pi} \int_{0}^{\pi} \frac{1}{x^{\gamma_{1}}y^{\gamma_{2}}} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} u_{i}(x)u_{j}(y) \right)^{p} dxdy \\ & < A(p,\gamma_{1},\gamma_{2}) \left(\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{m_{i}^{1-\frac{1}{p}(1-\gamma_{1})} s_{i}^{1-\frac{1}{p}(1-\gamma_{2})}}{n_{i}r_{j}} \right)^{p}. \end{split}$$

Finally, the use of the double version of the Levi's theorem implies the statement of the theorem.

It is clear that the conditions $\gamma_1, \gamma_2 > 1 - p$ in Theorem 2.3 are essential, therefore in the next theorem we examine the boundary case $\gamma_1, \gamma_2 = 1 - p$.

Theorem 2.4. If p = 2, 3, ... and $\gamma_1, \gamma_2 = 1 - p$, then the integral

$$\int_{0}^{\pi} \int_{0}^{\pi} \frac{G_{\Lambda_{1},\Lambda_{2}}^{p}(x, y)}{(xy)^{1-p}} dxdy$$

is convergent provided that the series

$$\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{1}{n_i r_j} (\log m_i) (\log s_j)$$

is convergent.

Proof. Observe first that in this boundary case the equality (2.8) reduces to the following:

(2.13)
$$\int_0^{\pi} \int_0^{\pi} \frac{1}{(xy)^{1-p}} \left(\sum_{i=1}^M \sum_{j=1}^N u_i(x) u_j(y) \right)^p dxdy$$

$$= \sum_{i_1=1}^M \cdots \sum_{i_p=1}^M \sum_{j_1=1}^N \cdots \sum_{j_p=1}^N \int_0^{\pi} \int_0^{\pi} \frac{1}{(xy)^{1-p}} u_{i_1}(x) \cdots u_{i_p}(x) u_{j_1}(y) \cdots u_{j_p}(y) dxdy.$$

Also, for $\gamma_1, \gamma_2 = 1 - p$ the estimates (2.9)-(2.12) take the following forms:

$$\int_{0}^{\alpha} \int_{0}^{\beta} \frac{1}{(xy)^{1-p}} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$\leq A^{2p} \int_{0}^{\alpha} \int_{0}^{\beta} \frac{1}{(xy)^{1-p}} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} dx dy = A^{2p} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{(\alpha \beta)^{p}}{p^{2}}$$

$$\int_{0}^{\alpha} \int_{\beta}^{\pi} \frac{1}{(xy)^{1-p}} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$\leq A^{2p} \int_{0}^{\alpha} \int_{\beta}^{\pi} \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{1}{r_{j_{1}} \cdots r_{j_{p}}} \frac{dx dy}{(xy)^{1-p}} \leq \frac{m_{i_{1}}}{n_{i_{1}}} \cdots \frac{m_{i_{p}}}{n_{i_{p}}} \frac{A^{2p}}{r_{j_{1}} \cdots r_{j_{p}}} \frac{\alpha^{1} \log \frac{\pi}{\beta}}{p}.$$

$$\int_{\alpha}^{\pi} \int_{0}^{\beta} \frac{1}{(xy)^{1-p}} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$\leq A^{2p} \int_{\alpha}^{\pi} \int_{0}^{\beta} \frac{1}{n_{i_{1}} \cdots n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{dx dy}{(xy)^{1-p}} \leq \frac{A^{2p}}{n_{i_{1}} \cdots n_{i_{p}}} \frac{s_{j_{1}}}{r_{j_{1}}} \cdots \frac{s_{j_{p}}}{r_{j_{p}}} \frac{\beta^{p} \log \frac{\pi}{n}}{p}$$

$$\int_{\alpha}^{\pi} \int_{t}^{\pi} \frac{1}{(xy)^{1-p}} u_{i_{1}}(x) \cdots u_{i_{p}}(x) u_{j_{1}}(y) \cdots u_{j_{p}}(y) dx dy$$

$$\leq A^{2p} \int_{\beta}^{\pi} \frac{1}{n_{i_{1}} \cdots n_{i_{p}}} \frac{1}{r_{j_{1}} \cdots r_{j_{p}}} \frac{dx dy}{(xy)^{1-p}} < \frac{A^{2p}}{n_{i_{1}} \cdots n_{i_{p}} r_{j_{1}} \cdots r_{j_{p}}} \log \frac{\pi}{\alpha} \log \frac{\pi}{\beta}.$$

respectively.

Next, specifying $\alpha=(m_{i_1}\cdots m_{i_p})^{-\frac{1}{p}}$ and $\beta=(s_{j_1}\cdots s_{j_p})^{-\frac{1}{p}}$ we obviously have

$$\log \frac{\pi}{\alpha} = \log \pi + \frac{1}{p} \log(m_{i_1} \cdots m_{i_p}) \quad \text{and} \quad \log \frac{\pi}{\beta} = \log \pi + \frac{1}{p} \log(s_{j_1} \cdots s_{j_p}).$$

Using these equalities, the above estimates and the equality (2.13)), we obtain

$$\int_{0}^{\pi} \int_{0}^{\pi} \frac{1}{|x_{1}|^{2} - n} \left(\sum_{i=1}^{M} \sum_{j=1}^{N} u_{i}(x) u_{j}(y) \right)^{p} dxdy$$

$$< \sum_{i_{1}=1}^{M} \cdots \sum_{i_{m}=1}^{M} \sum_{i_{m}=1}^{N} \cdots \sum_{i_{m}=1}^{M} \frac{1^{2p}}{n_{i_{1}} \cdots n_{i_{m}}} \left\{ \left(\frac{1}{p} + \log \pi \right)^{2} + \left(\frac{1}{p^{2}} + \log \pi \right) \left[\sum_{\nu=1}^{p} \log(m_{i_{\nu}}) + \sum_{\mu=1}^{p} \log(s_{j_{\mu}}) \right] + \frac{1}{p^{2}} \sum_{\nu=1}^{p} \sum_{\mu=1}^{p} \log(m_{i_{\nu}}) \log(s_{j_{\mu}}) \right\}.$$

Therefore, we have

$$\begin{split} & \int_{0}^{\pi} \int_{0}^{\pi} \frac{G_{\Lambda_{1},\Lambda_{2}}(x,y)}{(xy)^{-1}} dx dy \\ & \leq K A^{2p} \sum_{i_{1}=1}^{\infty} \cdot \sum_{i_{p}=1}^{\infty} \sum_{j_{1}=1}^{\infty} \cdot \sum_{j_{p}=1}^{\infty} \frac{1}{n_{i_{1}} \cdot n_{i_{p}} r_{j_{1}} \cdot \dots r_{j_{p}}} \sum_{\nu=1}^{p} \sum_{\mu=1}^{p} \log(m_{i_{\nu}}) \log(s_{j_{\mu}}) \\ & \leq K A^{2p} \left(\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{1}{n_{i} r_{j}} \right)^{p-1} \sum_{i=1}^{\infty} \sum_{j=1}^{\infty} \frac{\log(m_{i}) \log(s_{j})}{n_{i} r_{j}} \end{split}$$

where K is an absolute positive constant. The proof is completed.

Acknowledgment. The author would like to thank the anonymous referee for her/his remarks which improved the final form of this paper.

Список литературы

- S. A. Telyakovskii, "Certain properties of Fourier series of functions with bounded variation".
 East. J. Approx., 10, no. 1-2, 215 218 (2004).
- [2] S. A. Telyakovskii, "On the properties of blocks of terms of the series $\sum \frac{1}{k} \sin kx$ ", Ukrainian Math. J., 64, no. 5, 816 822 (2012).
- [3] R. M. Trigub, "A note on the paper of Telyakovskii "Certain properties of Fourier series of functions with bounded variation", East. J. Approx., 13, no. 1, 1 - 6 (2007).
- [4] Xh. Z. Krasniqi, "On a class of double trigonometric Fourier series of functions of bounded variation", East. J. Approx., 17, no. 4, 337 - 344 (2011).
- [5] F. Méricz, "Pointwise behavior of double Fourier series of functions of bounded variation", Monatsh. Math., 148, no. 1, 51 - 59 (2006)
- [6] V. P. Zastavnyi, "Estimates for sums of moduli of blocks from trigonometric Fourier series", Trudy Inst. Mat. i Mekh. UrO RAN, 16, no. 4, 166 – 179 (2010).

Поступила 12 февраля 2015

Известия НАП Армении, Математика, том 52, п. 1. 2017. стр. 68-77.

BRUCK CONJECTURE FOR A LINEAR DIFFERENTIAL POLYNOMIAL

1. LAHIRI AND B. PAL

University of Kalyani. India¹ E-mails: ilahiri@hatinail.com, palbipul86@qmail.com

Abstract. In the paper we study the Brück Conjecture for a linear differential polynomial.

MSC2010 numbers: 30D35.

Keywords: Meromorphic function. differential polynomial, sharing small function.

1. Introduction, definitions and results

Let f,g be nonconstant meromorphic functions defined in the open complex plane \mathbb{C} . For $a\in\mathbb{C}\cup\{\infty\}$ we say that f,g share the value a CM (counting multiplicities) if f,g have the same a-points with the same multiplicities, and we say that f,g share the value a IM (ignoring multiplicities) if f,g have the same a-points but the multiplicities are not taken into account.

The monograph [7] is a good source of standard notations and definitions of the value distribution theory. We now introduce some notation and a definition.

Definition 1.1. Given a meromorphic function f, a number $a \in \mathbb{C} \cup \{\infty\}$ and a positive integer k.

- (i) $N_{(k}(r,a;f)$ $(N_{(k}(r,a;f))$ denotes the counting function (reduced counting function) of those a-points of f whose multiplicatives are not less than k:
- (ii) $N_{k)}(r,a;f)$ $(\overline{N}_{k)}(r,a;f))$ denotes the counting function (reduced counting function) of those a-points of f whose multiplicatives are not greater than k;

Definition 1.2. A meromorphic function a = a(z) is called a small function of a meromorphic function f if T(r, a) = S(r, f).

In [5], R. Brück considered the uniqueness problem of an entire function when it shares a single value CM with its first derivative, and proposed the following conjecture, which inspired a number of people to work on the topic.

¹The work of the second author was supported by DAE (NBIIM fellowship), India.

Brück Conjecture: Let f be a nonconstant entire function satisfying $\nu(f) < \infty$, and let $\nu(f)$ be not a positive integer, where $\nu(f)$ is the hyper-order of f. If f and f' share one finite value $\sigma(f)$, then $f' - \sigma = c(f - \sigma)$ for some constant $c \neq 0$.

R. Brück [5] himself proved the following result

Theorem A ([5]). Let f be a nonconstant entire function. If f and f' share the value 1 CM and N(r, 0; f') = S(r, f), then f - 1 = c(f' - 1), where c is a nonzero constant.

Considering entire functions of finite order, L. Z. Yang [9] proved the following theorem.

Theorem B ([9]). Let f be a nonconstant entire function of finite order, and let $a(\neq 0)$ be a finite constant. If f and $f^{(k)}$ share the value a CM, then $f-a=c(f^{(k)}-a)$, where c is a nonzero constant and $k \geq 1$ is an integer.

In 2005, A. H. H. Al-khaladi [2] extended Theorem A to the class of meromorphic functions and proved the following result.

Theorem C ([2]). Let f be a nonconstant meromorphic function satisfying N(r,0;f') = S(r,f). If f and f' share the value 1 CM, then f-1 = c(f'-1) for some nonzero constant c.

Also, in [2] were considered the following examples, showing that the value sharing cannot be relaxed from CM to IM, and the condition N(r, 0; f') = S(r, f) is essential.

Example 1.1. Let $f = 1 + \tan z$. Then $f' - 1 = (f - 1)^2$ and $N(r, 0; f') \equiv 0$. Clearly f and f' share the value 1 IM but the conclusion of Theorem C does not hold.

Example 1.2. Let $f = \frac{1}{1+c^{-z}}$. Then f and f' share the value 1 CM and $N(r, 0; f') \neq S(r, f)$. It is easy to verify that $f' - 1 = \frac{1}{1+c^{z}}(f-1)$.

A. H. Al-khaladi [1] also observed by the following example that in Theorem A the shared value cannot be replaced by a shared small function.

Example 1.3. Let $f = 1 + e^e$ and $a = \frac{1}{1 - e^{-z}}$. Then a is a small function of f and f - a, f' - a share the value 0 CM and $N(r, 0; f') \equiv 0$. Also, we see that $f - a = \frac{1}{r}(f' - a)$

Considering the sharing of small functions, A. H. H. Al-khaladi [1] proved the following result.

Theorem D ([1]). Let f be a nonconstant entire function satisfying N(r,0;f')=S(r,f), and let $a=0,\infty$ be a meromorphic small function of f. If f-a and f'-a share the value 0 CM, then $f-a=\left(1+\frac{c}{a}\right)(f'-a)$, where $1+\frac{c}{a}=e^{\beta}$, c is a constant and β is an entire function.

For higher order derivatives, A. H. H. Al-khaladi [3] proved the following theorem.

Theorem E ([3]). Let f be a nonconstant entire function satisfying $N(r, 0; f^{(k)}) = S(r, f)$ (k > 1), and let $a \equiv 0, \infty$ be a meromorphic small function of f. If f - a and $f^{(k)} - a$ share the value 0 CM, then $f - a = \left(1 + \frac{1}{a}\right)(f^{(k)} - a)$, where P_{k-1} is a polynomial of degree at most k - 1 and $1 + \frac{1}{a} \not\equiv 0$.

Recently A. H. H. Al-khaladi [4] extended Theorem E to meromorphic functions. A natural extension of a derivative is a linear differential polynomial. For a transcendental meromorphic function f we denote by $L = L(f^{(k)})$ a linear differential polynomial of the form

(1.1)
$$L = L(f^{(k)}) = a_0 f^{(k)} + a_1 f^{(k+1)} + \dots + a_p f^{(k+p)},$$

where $a_0, a_1, \ldots, a_p \neq 0$ are constants, and $k \geq 1$ and $p \geq 0$ are integers such that p = 0 if k = 1 and $0 \leq p \leq k - 2$ if $k \geq 2$.

In the present paper we consider the problem of sharing a small function by a meromorphic function and a linear differential polynomial in conformity with Brück conjecture. The following theorem is the main result of the paper.

Theorem 1.1. Let f be a transcendental meromorphic function and let the differential polynomial $L = L(f^{(k)})$, given by (1.1), be nonconstant. Suppose that f - a and L - a share 0 CM, where $a \not\equiv 0, \infty$ is a small function of f. If $N(r, 0; f^{(k)}) = S(r, f)$, then

$$f-a=\left(1+\frac{P_{k-1}}{a}\right)(L-a),$$

where P_{k-1} is a polynomial of degree at most k-1 and $1+\frac{P_{k-1}}{a}\neq 0$.

The following example shows that the condition $N(r, 0; f^{(k)}) = S(r, f)$ is essential in Theorem 1.1.

Example 1.4. Let P be a nonconstant polynomial, and let $f = \frac{Pe^z}{1+e^z}$. Then $f' = \frac{e^z(P+P'+P'e^z)}{(1+e^z)^2}$ and hence $N(r,0;f') \neq S(r,f)$. Also, f-P' and f'-P' share 0. CM but $f'-P' = \frac{1}{1+e^z}(f-P')$, where T(r,P') = S(r,f).

2. Lemmas

In this section we present some necessary lemmas to be used in the proof of Theorem 1.1.

Lemma 2.1. Let f be a nonconstant meromorphic function and let $L = L(f^{(b)})$, given by (1.1), be nonconstant. If f - a and L - a share 0 CM, where $a = a(z) \not\equiv 0, \infty$ is a small function of f, then one of the following assertions holds:

(i)
$$f-a=\left(1+\frac{P_{k-1}}{a}\right)(L-a)$$
, where P_{k-1} is a polynomial of degree at most $k-1$ and $1+\frac{P_{k-1}}{a}\not\equiv 0$,

(ii)
$$T(r, f^{(k)}) \le (k + p + 1)\overline{N}(r, \infty; f) + \overline{N}(r, 0; f^{(k)}) + N(r, 0; f^{(k)}) + S(r, f)$$

Proof. Let $h = \frac{f-n}{L-a}$. Then h is an entire function and the poles of f are precisely the zeros of h. Now differentiating

$$(2.1) f - a = hL - ah$$

k-times we get

$$f^{(k)} - a^{(k)} = (hL)^{(k)} - (ha)^{(k)}.$$

We now consider the following cases.

CASE I. Let $a^{(k)} \not\equiv 0$. We put

(2.3)
$$W = \frac{(hL)^{(k)}}{hf^{(k)}} - \frac{(ha)^{(k)}}{ha^{(k)}}.$$

Since
$$W = \frac{(hL)^{(k)}}{hL} \cdot \frac{L}{f^{(k)}} - \frac{(ha)^{(k)}}{ha} \cdot \frac{a}{n^{(k)}}$$
, we have $m(r, W) = S(r, f)$.

We first suppose that $W \not\equiv 0$. Let z_0 be a zero of $f^{(k)} - a^{(k)}$ and $a^{(k)}(z_0) \neq 0, \infty$. Then from (2.2) we see that z_0 is a zero of $(hL)^{(k)} - (ha)^{(k)}$. Hence $W(z_0) = 0$ and we have

$$(2.4) N(r,0;f^{(k)} - a^{(k)}) \leq N(r,0;W) + S(r,f)$$

$$\leq T(r,W) + S(r,f)$$

$$= N(r,W) + S(r,f).$$

Also

$$(2.5) N(r,W) \le (k+p)N(r,\infty;f) + N(r,0;f^{(k)}) + S(r,f).$$

By Nevanlinna's three small functions theorem (see [7], p. 47), and formulas (2.4) and (2.5), we get

$$T(r,f^{(k)}) \le (k+p+1)\overline{N}(r,\infty;f) + \overline{N}(r,0;f^{(k)}) + N(r,0;f^{(k)}) + S(r,f),$$
 which is (ii).

Now let $W \equiv 0$. Then from (2.2) and (2.3) we get

$$(f^{(k)} - a^{(k)})a^{(k)} \equiv (ha)^{(k)}(f^{(k)} - a^{(k)}).$$

Since $f^{(k)} - a^{(k)} \not\equiv 0$, we obtain $(ha)^{(k)} \equiv a^{(k)}$. Integrating the last equality k-times we get $ha = a + P_{k-1}(z)$, where $P_{k-1}(z)$ is a polynomial of degree at most k-1. So $h = 1 + \frac{P_{k-1}}{a}$ and hence $f - a = \left(1 + \frac{P_{k-1}}{a}\right)(L-a)$, which is (i).

CASE II. Let $a^{(k)} \equiv 0$. Then a is a polynomial of degree at most k-1. From (2.2) we get $f^{(k)} = (hL)^{(k)} - (ah)^{(k)}$, and hence

(2.6)
$$\frac{1}{h} = \frac{(hL)^{(k)}}{hf^{(k)}} = \frac{(ah)^{(k)}}{hf^{(k)}}.$$

Putting $F = f^{(k)}$, $G = \frac{(hL)^{(k)}}{hf^{(k)}}$ and $b = \frac{(ah)^{(k)}}{h}$, from (2.6) we get

$$\frac{1}{h} = G - \frac{b}{F}$$

Differentiating (2.7) we obtain

(2.8)
$$-\frac{1}{h} \cdot \frac{h'}{h} = G' - \frac{b'}{F} + \frac{b}{F} \cdot \frac{F'}{F}$$

It follows from (2.7) and (2.8) that

$$\frac{A}{F} = G' + G \cdot \frac{h'}{h},$$

where $A = b \cdot \frac{h'}{h} + b' - b \cdot \frac{F'}{F}$

We first suppose that $G \equiv 0$. Then by integration we get $hL = Q_{k-1}$, where $Q_{k-1} = Q_{k-1}(z)$ is a polynomial of degree at most k-1. Putting $h = \frac{1-a}{L-a}$ we get

$$(2.10) (f-a)L = (L-a)Q_{k-1}.$$

Since a is a polynomial, from (2.10) we see that f is an entire function. Hence h is an entire function having no zeros. We put $h=e^{\alpha}$, where α is an entire function, and so $f=a+h(L-a)=a+Q_{k-1}-ae^{-\alpha}$ and $L=Q_{k-1}e^{-\alpha}$. It follows from the

definition of L that $L = R(a', \alpha')e^{\alpha}$, where $R(a', \alpha')$ is a differential polynomial in a'and α' . Hence

$$(2.11) R(\alpha', \alpha')e^{2\alpha} = Q_{k-1}$$

From (2.11) we see that $T(r,e^{\alpha}) = S(r,e^{\alpha})$, yielding a contradiction. Therefore $G \not\equiv 0$.

If h is a constant, say c, then f - a = c(L + a), which is (i).

Now we suppose that h is nonconstant and $b \equiv 0$. Then by integration we get $ah = P_{k-1}$, where $P_{k-1} = P_{k-1}(z)$ is a polynomial of degree at most k-1

Since h is an entire function and a is a polynomial of degree at most k-1, the equality $h = \frac{P_{k-1}}{a}$ implies that a is a factor of P_{k-1} , and hence

$$(2.12) h = Q_{k-t}^*,$$

where $Q_{k-t}^* = Q_{k-t}(z)$ is a polynomial of degree at most k-t $(t \ge 1)$.

If z_0 is a pole of f, then z_0 is a zero of h with multiplicity k + p, which is impossible by (2.12). So, f is an entire function, and hence h is an entire function having no zeros. Therefore from (2.12) we see that h is a constant, which is impossible.

Now we suppose that $b \not\equiv 0$. Let $A \equiv 0$, then from (2.9) we get $\frac{G'}{C} + \frac{h'}{h} \equiv 0$. By integration we obtain Gh = K and hence

$$(2.13) (hL)^{(k)} = Kf^{(k)},$$

where K is a nonzero constant. Again, $\frac{A}{h} = \frac{h}{h} + \frac{b}{h} - \frac{F}{E} = 0$ implies by integration hb = MF, and so

$$(2.14) (ah)^{(k)} = Mf^{(k)},$$

where M is a nonzero constant.

Since a is a polynomial and h is an entire function, we see from (2.14) that f is an entire function. So, h is an entire function having no zeros and we can put $h=e^{\alpha}$, where α is an entire function

Integrating (2.13) k-times we get

$$(2.15) hL = Kf + P_{k-1},$$

where $P_{k-1} = P_{k-1}(z)$ is a polynomial of degree at most k-1

Since hL = f - a + ah, from (2.15) we get

$$(2.16) (1-K)f = a(1-e^{\alpha}) + P_{k-1}^*.$$

If K=1, from (2.16) we see that $e^{\alpha}=1+\frac{P_{k-1}^{-}}{n}$, which is impossible. Hence $K\neq 1$. Now from (2.16) we get

$$f = \frac{ae^{\alpha}}{K-1} - \frac{a + P_{k-1}^*}{K-1}$$

Therefore from (1.1) we have

$$(2.18) L = R(\alpha')e^{\alpha},$$

where $R(\alpha')(\not\equiv 0)$ is a differential polynomial in α' with polynomial coefficients. From (2.15) we obtain

(2.19)
$$L = \frac{Ka}{K-1} - \frac{Ka + P_{k-1}^*}{K-1}e^{-\alpha}.$$

It follows from (2.18) and (2.19) that

$$K(\alpha')e^{2i\pi} = \frac{Ka}{K-1}c^{\alpha} - \frac{Ka + P_{k-1}}{K-1}.$$

This implies $T(r,e^{\alpha})=S(r,e^{\alpha})$, yielding a contradiction. Therefore $A\neq 0$.

Now observe that $A=b\left(\frac{h'}{h}+\frac{h'}{h}-\frac{F'}{F}\right)$ implies m(r,A)=S(r,f). Also, the poles of A are contributed by: (i) the poles of $b=\frac{(ah)^{(k)}}{h}$. (ii) the poles of $\frac{h'}{h}$ and (iii) the poles of $\frac{F'}{F}=\frac{f^{(k+1)}}{f^{(k)}}$. Since h is entire and the zeros of h are precisely the poles of f, and each zero of h is of multiplicity k+p, we get

$$N(r, A) \le (k+1)\overline{N}(r, \infty; f) + \overline{N}(r, 0; f^{(k)}) + S(r, f).$$

Therefore

$$(2.20) T(r,A) \le (k+1)\overline{N}(r,\infty;f) + \overline{N}(r,0;f^{(k)}) + S(r,f).$$

From (2.9) and (2.20) we get

$$m(r, \frac{1}{F}) \leq m(r, \frac{1}{A}) + m(r, G' + G\frac{h'}{h}) \leq T(r, A) + S(r, f)$$

$$\leq (k+1)N(r, \infty; f) + N(r, 0; f^{(k)}) + S(r, f).$$

So, by the first fundamental theorem, we obtain

$$T(r, f^{(k)}) \le (k+1)\overline{N}(r, \infty; f) + \overline{N}(r, 0; f^{(k)}) + N(r, 0; f^{(k)}) + S(r, f).$$

which implies (ii). This completes the proof of Lemma 2.1.

$$kN_{1)}(r,\infty;f) \leq \overline{N}_{(2}(r,\infty;f) + N_{1)}(r,\lambda;f^{(k)}) + \overline{N}(r,0;f^{(k+1)}) + S(r,f),$$

where \(\lambda\) is a constant.

Lemma 2.3 ([10], p.39). Let f be a nonconstant meromorphic function in the complex plane and let k be a positive integer. Then

$$N(r, 0; f^{(k)}) \le N(r, 0; f) + k\overline{N}(r, \infty; f) + S(r, f).$$

Lemma 2.4 ([8]). Given a transcendental meromorphic function f and a constant K > 1. Then there exists a set M(K) whose upper logarithmic density is at most

$$\delta(K) = \min\{(2e^{K-1} - 1)^{-1}, (1 + e(K - 1)\exp(e(1 - K)))\}$$

such that for every positive integer k.

$$\lim_{r \to \infty} \sup \frac{1}{T(r, |f^{(k)}|)} \le 3eK.$$

Lemma 2.5. Let f be a transcendental meromorphic function such that $N(r, 0; f^{(1)}) = S(r, f)$. If f + a and $a_1 f^{(1)} - a$ share 0 CM, where $a = a(z) (\not\equiv 0, \infty)$ is a small function of f and a_1 is a nonzero constant, then

$$N_{1)}(r, 0; f^{(2)}) \le \overline{N}_{(2}(r, \infty; f) + S(r, f).$$

Proof. If $a+a'\equiv 0$, then using the method of [4] (pp. 349 - 351), we get $N_{11}(r,0;f^{(2)})=S(r,f)$, and the result follows. If $a+a'\not\equiv 0$, then again using the method of [4] (pp. 351 - 354), we get $N_{11}(r,\infty;f)=S(r,f)$. Now by Lemma 2.3 we obtain

$$N(r, 0; f^{(2)}) \leq N(r, 0; f^{(1)}) + \overline{N}(r, \infty; f) + S(r, f)$$

= $\overline{N}_{(2)}(r, \infty; f) + S(r, f).$

Since $N_1(r, 0; f^{(2)}) \le N(r, 0; f^{(2)})$, the lemma is proved.

Lemma 2.6 ([6]). Let f be a transcendental meromorphic function and k be a positive integer. Then

$$k\overline{N}(r,\infty;f) \le N(r,0;f^{(k)}) + (1+\varepsilon)N(r,\infty;f) + S(r,f),$$

where e is any fixed positive number

3. PROOF OF THEOREM 1.1

Proof. First we verify that

(3.1)
$$\left(f^{(k+1)}\right)^{k+1} \neq c\left(f^{(k)}\right)^{k+2}$$
.

where $c \neq 0$ is a constant. Indeed, if (3.1) does not hold, then we get

(3.2)
$$\left(\frac{f^{(k+1)}}{f^{(k)}}\right)^{k+1} = cf^{(k)}.$$

Differentiating (3.2) and then using (3.2) we obtain

$$\left(\frac{f^{(k+1)}}{f^{(k)}}\right)^{-2} \left(\frac{f^{(k+1)}}{f^{(k)}}\right)' = \frac{1}{k+1}$$

Integrating twice we get

$$f^{(k)} = \frac{1}{\{Cz + D(k+1)\}^{k+1}},$$

where $C \neq 0$ and D are constants. This is impossible because f is transcendental. Let $k \geq 2$. We suppose that

$$T(r,f^{(k)}) \leq (k+p+1)\overline{N}(r,\infty;f) + \overline{N}(r,0;f^{(k)}) + N(r,0;f^{(k)}) + S(r,f).$$

Since $N(r, 0; f^{(k)}) = S(r, f)$, we get from above

(3.3)
$$T(r, f^{(k)}) \le (k + p + 1)\overline{N}(r, \infty; f) + S(r, f).$$

Also, from Lemma 2.6 we obtain for $0 < \varepsilon < \frac{k}{p+1} - 1$,

$$kN(r,\infty;f) \le (1+\epsilon)N(r,\infty;f) + S(r,f).$$

Hence from (3.3) we obtain

$$m(\tau, f^{(k)}) + N(\tau, \infty; f) \le \frac{p+1}{k} (1+\varepsilon)N(\tau, \infty; f) + S(\tau, f)$$

and so $m(r, f^{(k)}) + N(r, \infty; f) = S(r, f)$. Therefore

(3.4)
$$T(r, f^{(k)}) = S(r, f).$$

Let M(K) be defined as in Lemma 2.4. By (3.4) we can choose a sequence $r_n \to \infty$ such that $r_n \notin M(K)$ and $\lim_{n \to \infty} \frac{T(r_n, f^{(k)})}{T(r_n, f)} = 0$. This contradicts Lemma 2.4. Next, let k = 1. We suppose

$$T(r, f^{(1)}) \le 2\overline{N}(r, \infty; f) + \overline{N}(r, 0; f^{(1)}) + N(r, 0; f^{(1)}) + S(r, f).$$

Since $N(r, 0; f^{(1)}) = S(r, f)$, we obtain

$$m(r, f^{(1)}) + N(r, \infty; f) \le \overline{N}(r, \infty; f) + S(r, f)$$

and so

(3.5)
$$m(r, f^{(1)}) + N_{i2}(r, \infty; f) = S(r, f).$$

By the second fundamental theorem we get in view of (3.5)

$$T(r,f^{(1)}) \leq N(r,1;f^{(1)}) + N(r,0;f^{(1)}) + \overline{N}(r,\infty;f) - N(r,0;f^{(2)}) + S(r,f)$$

and so

(3.6)
$$m(r,1;f^{(1)}) + N(r,0;f^{(2)}) \le N_{1}(r,\infty;f) + S(r,f).$$

Now by Lemma 2.2 and (3.5) we get for $\lambda = 0$

(3.7)
$$N_{1}(r,\infty;f) \le \overline{N}(r,0,f^{(2)}) + S(r,f)$$

From (3.6) and (3.7) we get

(3.8)
$$N_{(2}(r,0;f^{(2)}) = S(r,f).$$

By (3.5), (3.8) and Lemma 2.5 we obtain

(3.9)
$$N(r,0;f^{(2)}) = S(r,f)$$

Hence by (3.5), (3.7) and (3.9) we get $N(r, \infty; f) = S(r, f)$, and so by (3.5) we have $T(r, f^{(1)}) = S(r, f)$ which is (3.4) for k = 1. Similarly using Lemma 2.4 we arrive at a contradiction. Therefore by Lemma 2.1 we obtain

$$f - a = \left(1 + \frac{P_{k-1}}{a}\right)(L - a).$$

-Список литературы

- A. H. H. Al-khaladi, "On entire functions which share one small function CM with their first derivative", Kodai Math. J., 27, 201 - 205 (2004).
- A. H. H. Al-khaladi. "On meromorphic functions that share one value with their derivative". Analysis, 25, 131 - 140 (2005).
- A. H. H. Al-khaladi, "On entire functions which share one small function CM with their kth derivative", Results Math., 47, 1 5 (2005).
- A. H. H. Al-khaladi, "Meromorphic functions that share one small function with their kth derivative", Analysis, 31, 341 354 (2011).
- R. Bruck, "On entire functions share one value CM with their first derivative", Results Math., 30, 21 – 24 (1996).
- G. Frang and G. Weissenborn, "On the zeros of linear differential polynomials of meromorphic functions", Complex Variables, 12, 77 – 81 (1989).
- 7. W. K. Hayman, Meromorphic Functions, The Clarendon Press, Oxford (1964).
- W. K. Hayman and J. Miles, "On the growth of a meromorphic function and its derivatives", Complex Variables, 12, 245 - 260 (1989).
- L. Z. Yang, "Solution of a differential equation and its applications", Kodai Math. J., 22(3), 458

 464 (1999).
- C. C. Yang and H. X. Yi, Uniqueness Theory of Meromorphic Functions, Science Press and Kluwer Academic Publishers (2003).

Поступила 27 марта 2015

Известия ПАП Армении, Математика, том 52, п. 1. 2017, стр. 78-84.

FIXED POINTS OF MIXED MONOTONE OPERATORS FOR EXISTENCE AND UNIQUENESS OF NONLINEAR FRACTIONAL DIFFERENTIAL EQUATIONS

H. R. MARASI, H. AFSHARI, M. DANESHBASTAM, C.B. ZHAI

University of Tabriz, Tabriz, Iran
University of Bonab, Bonab, Iran
Shanxi University, Taiyuan, Shanxi, China
E-mails: marasi@bonabu.ac.ir.; hojat.afshari@yahoo.com
daneshbastam79@gmail.com; cbzhai@sxu.edu.cn

Abstract. In this paper we study the existence and uniqueness of positive solutions for nonlinear fractional differential equation boundary value problems by using new fixed point results of mixed monotone operators on cones.

MSC2010 numbers: 30D35.

Keywords: Fractional differential equation; normal cone; boundary value problem; mixed monotone operator.

1. Introduction

In recent years, boundary value problems for nonlinear fractional differential equations with a variety of boundary couditions have been investigated by many researchers. Fractional differential equations appear naturally in various fields of science and engineering, and thus constitute an important field of research (see [1 3]). As a matter of fact, fractional derivatives provide a powerful tool for the description of memory and hereditary properties of various materials and processes. A significant feature of a fractional order differential operator, in contrast to its counterpart in classical calculus, is its nonlocal behavior, meaning that the future state of a dynamical system or process based on the fractional differential operator depends on its current state as well its past states. In other words, differential equations of arbitrary order are capable of describing memory and hereditary properties of certain important materials and processes. This aspect of fractional calculus has contributed towards the growing popularity of the subject. Mixed monotone operators were introduced by Guo and Lakshmikantham in [4]. Their study has wide applications in the applied sciences such as engineering, biological chemistry technology, nuclear physics and in mathematics (see [6 8]). Various existence and uniqueness theorems of fixed points for mixed monotone operators have been obtained by a number of authors (see [9] [12]). Bhaskar and Lakshmikantham, [9], established some coupled fixed point theorems for mixed monotone operators in partially ordered metric spaces and discussed a question of existence and uniqueness of a solution for a periodic boundary value problem. Recently Y. Sang. [13], proved some new existence and uniqueness theorems of a fixed point of mixed monotone operators with perturbations.

In this paper, by applying Sang's results, we obtain some new results on the existence and uniqueness of positive solutions for some nonlinear fractional differential equations via given boundary value problems.

We first introduce some notations, definitions and known results to be used in the paper.

Definition 1.1 ([1, 2]). For a continuous function $f:[0,\infty)\to\mathbb{R}$, the Caputo derivative of fractional order α is defined by

$${}^{\circ}D^{\alpha}f(t) = \frac{1}{\Gamma(n-\alpha)} \int_0^t (t-s)^{n-\alpha-1} f^{(n)}(s) ds.$$

where $n-1 < \alpha < [\alpha] + 1$ and $[\alpha]$ denotes the integer part of α .

Definition 1.2 ([1, 2]). The Riemann-Liouville fractional derivative of order α for a continuous function f is defined by

$$D^{\alpha}f(t) = \frac{1}{\Gamma(n-\alpha)} \left(\frac{d}{dt}\right)^n \int_0^t \frac{f(s)}{(t-s)^{\alpha-n-1}} ds, \qquad n = [\alpha] + 1.$$

where the right-hand side is defined pointwise on $(0, \infty)$.

Definition 1.3 ([1, 2]). Let [a,b] be an interval in \mathbb{R} and $\alpha > 0$. The Riemann-Liouville fractional order integral of a function $f \in L^1([a,b],\mathbb{R})$ is defined by

$$I_a^{\alpha} f(t) = \frac{1}{\gamma(\alpha)} \int_a^t \frac{f(s)}{(t-s)^{1+\alpha}} ds.$$

whenever the integral exists.

Let $(E, \|\cdot\|)$ be a Banach space which is partially ordered by a cone $P \subseteq E$, that is, $x \leq y$ if and only if $y - x \in P$. If $x \neq y$, then we denote x < y or x > y. Also, the zero element of E we denote by θ . Recall that a non-empty closed convex set $P \subset E$ is called a cone if it satisfies the conditions: (i) $x \in P$, $\lambda \geq 0 \Longrightarrow \lambda x \in P$. (ii) $x \in P$, $-x \in P \Longrightarrow x = \theta$. A cone P is called normal if there exists a constant N > 0 such that $\theta \leq x \leq y$ implies $\|x\| \leq N \|y\|$. Also, we define the order interval $[x_1, x_2] = \{x \in E | x_1 \leq x \leq x_2\}$ for all $x_1, x_2 \in E$. We say that an operator $A : E \to E$ is increasing whenever $x \leq y$ implies $Ax \leq Ay$.

Definition 1.4 ([4, 5]). Let $D \subset E$. An operator $A: D \times D \to D$ is said to be a mixed monotone operator if A(x,y) is increasing in x and decreasing in y, that is, $u_i, v_i \in D$ $(i = 1, 2), u_1 \leq u_2, v_1 \geq v_2$ implies $A(u_1, v_1) \leq A(u_2, v_2)$.

An element $x^* \in D$ is called a fixed point of A if it satisfies $A(x^*, x^*) = x^*$. For $h > \theta$ we define $P_h = \{x \in E | \exists \lambda, \mu > 0; \ \lambda h \le x \le \mu h\}$.

In this paper, using the existence and uniqueness results for the solution of the following operator equation

$$(1.1) A(x,x) + Bx = x,$$

where A is a mixed monotone operator, B is sublinear and E is a real ordered Banach space, obtained in [13] by the partial ordering theory and monotone iterative technique, we study a question of existence and uniqueness of positive solutions for nonlinear fractional differential equation boundary value problems.

Theorem 1.1 ([13]). Let P be a normal cone in E. $A: P \times P \rightarrow P$ be a mixed monotone operator, and let $B: E \rightarrow E$ be sublinear. Assume that for all a < t < b, there exist two positive-valued functions $\tau(t)$ and $\varphi(t, x, y)$ defined on an interval (a, b) such that:

- (II_1) $\tau:(a,b) \rightarrow (0,1)$ is surjection;
- (H_2) $\varphi(t, x, y) > \tau(t)$ for all $t \in (a, b), x, y \in P$;
- (H_3) $\Lambda(\tau(t)x, \frac{1}{\tau(t)}y) \ge \varphi(t, x, y)\Lambda(x, y)$ for all $t \in (a, b), x, y \in P$:
- (H_4) $(I-B)^{-1}: E \to E$ exists and is an increasing operator.

Furthermore, for any $t \in (a,b)$ the function $\phi(t,x,y)$ is nonincreasing in x for fixed y, and nondecreasing in y for fixed x. In addition, suppose that there exist $h \in P - \{\theta\}$ and $t_0 \in (a,b)$ such that

$$\tau(t_0)h \le (I - B)^{-1}A(h, h) \le \frac{\pi(h_0 - h_0)}{\pi(h_0)}h.$$

Then the following assertions hold:

(i) there are $u_0, v_0 \in P_h$ and $r \in (0, 1)$ such that $rv_0 \le u_0 \le v_0$ and

$$u_0 \le (I - B)^{-1} \Lambda(u_0, v_0) \le (I - B)^{-1} \Lambda(v_0, u_0) \le v_0;$$

- (ii) the equation (1.1) has a unique solution x^* in $[u_0, v_0]$;
- (iii) for any initial values $x_0, y_0 \in P_h$, constructing successively the sequences

$$x_n = (I - B)^{-1} A(x_{n-1}, y_{n-1}), \quad y_n = (I - B)^{-1} A(y_{n-1}, x_{n-1}), \quad n = 1, 2, \dots$$

we have $||x_n - x^*|| \to 0$ and $||y_n - x^*|| \to 0$ as $n \to \infty$.

2. MAIN RESULTS

We study the existence and uniqueness of a solution of a fractional differential equation on a partially ordered Banach space with two types of boundary conditions and two types of fractional derivatives. We first study the existence and uniqueness of a positive solution for the following fractional differential equation:

(2.1)
$$\frac{D^{\alpha}}{Dt}u(t) = f(t, u(t), u(t)), \qquad t \in [0, 1], \ 3 < \alpha \le 4,$$

subject to conditions

(2.2)
$$u(0) = u'(0) = u(1) = u'(1) = 0,$$

where D^{α} is the Riemann-Liouville fractional derivative of order α .

Consider the Banach space of continuous functions on [0,1] with sup norm and set $P = \{y \in C[0,1] : \min_{t \in [0,1]} y(t) \ge 0\}$. Then P is a normal cone. The next two lemmas were proved in [14].

Lemma 2.1 ([14]). Given $y \in C[0,1]$ and a number α such that $3 < \alpha \le 4$. Then the unique solution of the following fractional differential equation boundary value problem.

(2.3)
$$\frac{D^{\alpha}}{Dt}u(t) = f(t, y(t)), t \in [0, 1], \ 3 < \alpha \le 4,$$
$$u(0) = u'(0) = u(1) = u'(1) = 0,$$

is given by

$$u(t) = \int_0^1 G(t, s) f(s, y(s)) ds,$$

where

$$(2.4) G(t,s) = \begin{cases} \frac{(t-1)^{\alpha-1}+(1-s)^{\alpha-2}t^{\alpha-2}|(s-t)+(\alpha-2)(1-t)s|}{1+\alpha}, & 0 \le s \le t \le 1, \\ \frac{(1-s)^{\alpha-2}t^{\alpha-2}[(s-t)+(\alpha-2)(1-t)s]}{\Gamma(\alpha)}, & 0 \le t \le s \le 1. \end{cases}$$

If f(t, u(t)) = 1, then the unique solution of (2.3) is given by $u_0(t) = \int_0^t G(t, s) ds = \frac{1}{\Gamma(n+1)} t^{\alpha-2} (1-t)^2$.

Lemma 2.2 ([14]). The Green's function G(t,s) has the following properties:

(1) G(t,s) > 0 and G(t,s) is continuous for $t,s \in [0,1]$:

(2)
$$\frac{(\alpha-2)h(t)k(s)}{\Gamma(\alpha)} \le G(t,s) \le \frac{M_0k(s)}{\Gamma(\alpha)},$$
where $M_0 = \max\{\alpha - 1, (\alpha - 2)^2\}$. $h(t) = t^{\alpha-2}(1-t)^2$, $k(s) = s^2(1-s)^{\alpha-2}$.

Now we are ready to state and prove our first main result.

Theorem 2.1. Let $f(t, u(t), v(t)) \in C([0, 1] \times [0, \infty) \times [0, \infty))$ be an increasing in u and decreasing in v function. Assume that for all a < t < b there exist two positive-valued functions $\tau(t)$ and $\varphi(t, u, v)$ defined on an interval (a, b) such that:

$$(H_1)$$
 $\tau:(a,b)\to(0,1)$ is surjection;

$$(H_2) \varphi(t, u, v) > \tau(t)$$
 for all $t \in (a, b), u, v \in P$;

$$(H_3) \int_0^1 G(t,s) f(s,\tau(t)u(s), \frac{1}{\tau(t)}v(s)) \ge \varphi(t,u,v) \int_0^1 G(t,s) f(s,u(s),v(s)) ds.$$

Furthermore, for any $t \in (a,b)$ the function $\varphi(t,u,v)$ is nonincreasing in u for fixed v, and nondecreasing in v for fixed u. In addition, suppose that there exist $h \in P - \{\theta\}$

An element $x^* \in D$ is called a fixed point of A if it satisfies $A(x^*, x^*) = x^*$. For $h > \theta$ we define $P_h = \{x \in E | \exists \lambda, \mu > 0; \ \lambda h \leq x \leq \mu h\}$.

In this paper, using the existence and uniqueness results for the solution of the following operator equation

$$(1.1) A(x,x) + Bx = x.$$

where A is a mixed monotone operator, B is sublinear and E is a real ordered Banach space, obtained in [13] by the partial ordering theory and monotone iterative technique, we study a question of existence and uniqueness of positive solutions for nonlinear fractional differential equation boundary value problems.

Theorem 1.1 ([13]). Let P be a normal cone in E, $A: P \times P \to P$ be a mixed monotone operator, and let $B: E \to E$ be sublinear. Assume that for all a < t < b, there exist two positive-valued functions $\tau(t)$ and $\varphi(t, x, y)$ defined on an interval (a, b) such that:

- (H_1) $\tau:(a,b)\to(0,1)$ is surjection;
- (H_2) $\varphi(t,x,y) > \tau(t)$ for all $t \in (a,b), x,y \in P$;
- (H_3) $A(\tau(t)x, \frac{1}{\tau(t)}y) \ge \varphi(t, x, y)A(x, y)$ for all $t \in (a, b), x, y \in P$:
- (H_4) $(I-B)^{-1}: E \to E$ exists and is an increasing operator.

Furthermore, for any $t \in (a,b)$ the function $\phi(t,x,y)$ is nonincreasing in x for fixed y, and nondecreasing in y for fixed x. In addition, suppose that there exist $h \in P - \{\theta\}$ and $t_0 \in (a,b)$ such that

$$\tau(t_0)h \le (I-B)^{-1}A(h,h) \le \frac{\varphi(t_0,\frac{h}{\tau(t_0)},-(t_0)h)}{\tau(t_0)}h.$$

Then the following assertions hold:

(i) there are $u_0, v_0 \in P_h$ and $r \in (0, 1)$ such that $rv_0 \le u_0 \le v_0$ and

$$u_0 \le (I-B)^{-1} A(u_0, v_0) \le (I-B)^{-1} A(v_0, u_0) \le v_0;$$

- (ii) the equation (1.1) has a unique solution x^* in $[u_0, v_0]$;
- (iii) for any initial values $x_0, y_0 \in P_h$, constructing successively the sequences

$$x_n = (I - B)^{-1} A(x_{n-1}, y_{n-1}), \quad y_n = (I - B)^{-1} A(y_{n-1}, x_{n-1}), \quad n = 1, 2, \dots,$$

we have $||x_n - x^*|| \to 0$ and $||y_n - x^*|| \to 0$ as $n \to \infty$.

2. MAIN RESULTS

We study the existence and uniqueness of a solution of a fractional differential equation on a partially ordered Banach space with two types of boundary conditions and two types of fractional derivatives. We first study the existence and uniqueness of a positive solution for the following fractional differential equation:

(2.1)
$$\frac{D^{\alpha}}{Dt}u(t) = f(t, u(t), u(t)), \qquad t \in [0, 1], \ 3 < \alpha \le 4,$$

subject to conditions

(2.2)
$$u(0) = u'(0) = u(1) = u'(1) = 0,$$

where D^{α} is the Riemann-Llouville fractional derivative of order α .

Consider the Banach space of continuous functions on [0, 1] with sup norm and set $P = \{y \in C[0,1] : \min_{t \in [0,1]} y(t) \ge 0\}$. Then P is a normal cone. The next two lemmas were proved in [14].

Lemma 2.1 ([14]). Given $y \in C[0,1]$ and a number α such that $3 < \alpha \le 4$. Then the unique solution of the following fractional differential equation boundary value problem

(2.3)
$$\frac{D^{\alpha}}{Dt}u(t) = f(t, y(t)), \qquad t \in [0, 1], \ 3 < \alpha \le 4.$$
$$u(0) = u'(0) = u(1) = u'(1) = 0,$$

is given by

$$u(t) = \int_0^1 G(t, s) f(s, y(s)) ds,$$

where

$$(2.4) G(t,s) = \begin{cases} \frac{(t-1)^{n-s} + (1-s)^{n-2} t^{n-2} (s-t) + (\alpha-2)(1-t)s!}{1(\alpha)}, & 0 \le s \le t \le 1, \\ \frac{(1-s)^{n-2} + n^{n-2} ((s-t) + (\alpha-2)(1-t)s!}{1(\alpha)}, & 0 \le t \le s \le 1. \end{cases}$$

If f(t, u(t)) = 1, then the unique solution of (2.3) is given by $u_0(t) = \int_0^1 G(t, s) ds = \frac{1}{\Gamma(\alpha+1)} t^{\alpha-2} (1-t)^2.$

Lemma 2.2 ([14]). The Green's function G(t,s) has the following properties.

(1)
$$G(t,s) > 0$$
 and $G(t,s)$ is continuous for $t,s \in [0,1]$:
(2)
$$\frac{(\alpha-1)h(t)h(s)}{(\alpha)} \le G(t,s) \le \frac{M_0k(s)}{\Gamma(\alpha)},$$
where $M_0 = \max\{\alpha - 1, (\alpha - 2)^2\}, \quad h(t) = t^{\alpha-2}(1-t)^2, \quad k(s) = s^2(1-s)^{\alpha-2}$

Now we are ready to state and prove our first main result.

Theorem 2.1. Let $f(t, u(t), v(t)) \in C([0, 1] \times [0, \infty) \times [0, \infty))$ be an increasing in uand decreasing in v function. Assume that for all a < t < b there exist two positivevalued functions $\tau(t)$ and $\varphi(t, u, v)$ defined on an interval (a, b) such that:

 (H_1) τ , $(a,b) \rightarrow (0,1)$ is surjection;

 (H_2) $\varphi(t, u, v) > \tau(t)$ for all $t \in (a, b), u, v \in P$;

 $(H_3) \int_0^1 G(t,s) f(s,\tau(t)u(s), \tfrac{1}{\tau(t)}v(s)) \ge \varphi(t,u,v) \int_0^1 G(t,s) f(s,u(s),v(s)) ds.$

Furthermore, for any $t \in (a,b)$ the function $\varphi(t,u,v)$ is nonincreasing in u for fixed v_i and nondecreasing in v for fixed u. In addition, suppose that there exist $h \in P - \{\theta\}$

and $t_0 \in (a,b)$ such that

(2.5)
$$\tau(t_0)h \leq \int_0^1 G(t,s)f(s,h(s),h(s))ds \leq \frac{\tau(t_0)}{\tau(t_0)}, \tau(t_0)h.$$

Then the following assertions hold:

- (i) there are $u_0, v_0 \in P_h$ and $r \in (0, 1)$ such that $rv_0 \le u_0 \le v_0$, and $u_0 \le \int_0^1 G(t, s) f(s, u_0(s), v_0(s)) ds \le \int_0^1 G(t, s) f(s, v_0(s), u_0(s)) ds \le v_0$,
- (ii) the problem (2.1), (2.2) has a unique solution x^* in $[u_0, v_0]$,
- (iii) for any initial values $u_0, v_0 \in P_h$ and $n = 1, 2, \dots$, constructing successively the sequences

$$u_n = \int_0^1 G(t,s) f(s,u_{n-1}(s),v_{n-1}(s)) ds, \quad v_n = \int_0^1 G(t,s) f(s,v_{n-1}(s),u_{n-1}(s)) ds,$$
we have $||u_n - u^*|| \to 0$ and $||v_n - v^*|| \to 0$ as $n \to \infty$.

Proof. By Lemma 2.1, the problem is equivalent to equation $u(t) = \int_0^1 G(t,s) f(s,y(s)) ds$, where G(t,s) is defined by (2.4). Define the operator $A: P \times P \to E$ as follows: $A(u(t),v(t)) = \int_0^1 G(t,s) f(s,u(s),v(s)) ds$, and observe that u is a solution for the problem if and only if u = A(u,u).

Next, it is easy to see that the operator A is increasing in u and decreasing in v on P. Hence, under the assumptions of the theorem we have $A(\tau(t)u, \frac{1}{\tau(t)}v) \geq \varphi(t, u, v)A(u, v)$ for all $t \in (a, b)$, $u, v \in P$ and $\tau(t_0)h \leq A(h, h) \leq \frac{1}{\tau(t)}h$. Thus, the operator A satisfies all the conditions of Theorem 1.1, and hence A has a unique positive solution (u^*, u^*) such that $A(u^*, u^*) = u^*$, $u^* \in [u_0, v_0]$.

Example 2.1. Consider the following periodic boundary value problem:

(2.6)
$$D^{\frac{7}{2}}u(t) = f(t, u(t), u(t)) = g(t) + u(t) + \frac{1}{\sqrt{u(t)}}. \qquad t \in [0, 1].$$
$$u(0) = u'(0) = u(1) = u'(1) = 0,$$

where g(t) is continuous on [0,1] with $388.625 \le g(t) \le 63728$.

For every $\lambda \in (0,1)$ and $u,v \in P$ we have $\int_0^1 G(t,s)[q(s) + \lambda u(s) + \frac{1}{\sqrt{\frac{1}{\lambda} v(s)}}]ds = \lambda \int_0^1 G(t,s)[\frac{g(s)}{\lambda} + u(s) + \frac{1}{\sqrt{\lambda v(s)}}]ds$ $\leq \lambda \frac{\frac{g(s)}{\lambda} + u(s) + \frac{1}{\sqrt{\lambda v(s)}}}{g(s) + u(s) + \frac{1}{\sqrt{\lambda v(s)}}} \int_0^1 G(t,s)[g(s) + u(s) + \frac{1}{\sqrt{\lambda v(s)}}]ds$

We note that

$$\lambda < \varphi(\lambda,u,v) = \lambda \frac{1}{g(s)+u(s)+\sqrt{\log s}} \int_0^1 G(t,s)[g(s)+u(s)+\frac{1}{\sqrt{v(s)}}] ds < 1.$$

By means of some calculations, we can conclude that for any $\lambda \in (0,1)$ the function φ is nonincreasing in u for fixed v and nondecreasing in v for fixed u.

So, it is enough to verify that the condition (2.6) of Theorem 2.1 is satisfied. Putting u=v=h=1, and taking into account that $M_1=\min_{t\in[0,1]}\int_{\Omega}G(t,s)ds=0.00001$ and $M_2 = \max_{t \in [0,1]} \int_0^t G(t,s) ds = 0.004$, we can easily get $2^{-8} \le 0.00001 \times 390.625 \le \int_0^1 G(t,s)[g(s) + u(s) + \frac{1}{\sqrt{s}}]ds$

implying that the condition (2.5) of Theorem 2.1 holds. Therefore, we can apply this theorem to conclude that the problem in the example has a unique solution.

Now, we study the existence and uniqueness of a positive solution for the following fractional differential equation:

(2.7)
$${}^{c}D^{\alpha}y(t) = h(t), \quad t \in [0, T], \quad T \ge 1.$$

subject to

(2.8)
$$y(0) + \int_0^T y(s)ds = y(T).$$

Lemma 2.3 ([15]). Let $0 < \alpha \le 1$ and let $h \in C([0,T],\mathbb{R})$ be a given function. Then the boundary value problem (2.7), (2.8) has a unique solution given by

$$y(t) = \int_0^T G(t, s)h(s)ds,$$

$$G(t,s) = \begin{cases} \frac{-(T-s)^{\alpha} + \alpha T(t-s)^{\alpha}}{T\Gamma(\alpha+1)} + \frac{(T-s)^{\alpha}}{T\Gamma(\alpha)}, & 0 \le c < t, \\ \frac{(T-s)^{\alpha}}{T\Gamma(\alpha+1)} + \frac{(T-s)^{\alpha-1}}{T\Gamma(\alpha)}, & t \le s < T. \end{cases}$$

By using arguments similar to those applied in the proof of Theorem 2.1, it can easily be verified that Theorem 2.1 remains true for Green function defined in Lemma 2.3.

Example 2.2. Consider the following boundary value problem

$$^{*}D^{\frac{1}{2}}u(t) = f(t, u(t), u(t)) = g(t)u(t)^{\frac{1}{2}} + u(t)^{\frac{-1}{4}}, \qquad t \in [0, 1],$$

$$u(0) + \int_{0}^{1} u(s)ds = u(1).$$

where g(t) is continuous on [0,1] with $0.6378 \le g(t) \le 2.89967$.

For every $\lambda \in (0,1)$ and $u, v \in P$ we have

$$(2.9) \qquad \int_{0}^{1} G(t,s)[g(s)((\lambda u(s))^{\frac{1}{2}} + (\frac{1}{\lambda}v(s))^{\frac{-1}{3}})]ds \ge \lambda \int_{0}^{1} G(t,s)[\frac{g(s)}{\lambda} + u(s) + \frac{1}{\sqrt{v(s)}}]ds \ge \lambda \int_{0}^{1} G(t,s)[(g(s)(\frac{u(s)}{\lambda})^{\frac{1}{2}} + (\lambda v(s))^{\frac{-1}{3}})]ds$$

$$\ge \lambda \frac{g(s)(\frac{u(s)}{\lambda})^{\frac{1}{2}} + (\lambda v(s))^{\frac{-1}{3}}}{g(s)u(s)^{\frac{1}{2}} + v(s)^{\frac{-1}{3}}} \int_{0}^{1} G(t,s)[g(s)u(s)^{\frac{1}{2}} + v(s)^{\frac{-1}{3}}]ds.$$

Note that

$$\lambda < \varphi(\lambda, u, v) = \lambda \frac{g(s)(\frac{u(s)}{3})^{\frac{1}{s}} + (\lambda v(s))^{\frac{-1}{3}}}{g(s)u(s)^{\frac{1}{2}} + v(s)^{\frac{-1}{3}}} < 1.$$

By means of some calculations, we can conclude that for any $\lambda \in (0,1)$ the function φ is nonincreasing in u for fixed v and nondecreasing in v for fixed u.

So, it is enough to verify that the condition (2.5) of Theorem 2.1 is satisfied. Putting u=v=h=1, and taking into account that $M_1=\min_{t\in [0,1]}\int_0^1 G(t,s)ds=\frac{1}{2}$, and $M_2=\max_{t\in [0,1]}\int_0^1 G(t,s)ds=\frac{80}{51}$, we can easily get $2^{-6}\leq \frac{1}{4}\times 1.6378\leq \int_0^1 G(t,s)[g(s)u(s)^{\frac{n}{2}}+u(s)^{\frac{n-1}{2}}]ds\leq \frac{80}{51}\times 3.8967 \leq \frac{16g(s)+8}{2g(s)+1}=\frac{1}{2}$ $s\in [0,1]$, implying that the condition (2.5) of Theorem 2.1 holds. Therefore, we can apply this theorem to conclude that the problem in the example has a unique solution.

Список литературы

- [1] I Podlubny, Fractional Differential Equations, Academic Press, San Diego (1999).
- A. A. Kilbas, H. M. Srivastava, J. J. Trujillo, "Theory and applications of fractional differential equations, North-Holland mathematics studies", 204.7 - 10 (2006).
- [3] J. Sabatier, O. P. Agrawal, J. A. T. Machado, Advances in Fractional Calculus: Theoretical Developments and Applications in Physics and Engineering, Springer, Dordrecht (2007).
- [4] D. Guo, V. Lakskmikantham, "Coupled fixed points of nonlinear operators with applications", Nonlinear Anal. 11 (5), 623 - 632 (1987).
- [5] D. Guo, "Fixed points of mixed monotone operators with application", Appl. Anal. 34, 215 224 (1988).
- [6] D. Guo, V. Lakskmikantham, Nonlinear Problems in Abstract Cones, Academic Press, New York (1988).
- [7] D. Guo, Partial Order Methods in Nonlinear Analysis, Jinan, Shandong Science and Technology Press [in Chinese] (2000).
- [8] S. S. Zhang, "Coupled fixed points for mixed monotone condensing operators and an existence theorem of the solution for a class of functional equations arising in dynamic programming", J. Math. Appl. 160, 468 479 (1991).
- [9] T. G. Bhaskar, V. Lakshmikambam, "Fixed point theorems in partially ordered metric spaces and applications", Nonlinear Anal. TMA 65, 1379 - 1393 (2006).
- [10] Dz. Burgic, S. Kalabusic, M. R. S. Kulenovic, "Global attractivity results for mixed monotone mappings in partially ordered complete metric spaces", Fixed Point Theory Appl. Article ID 762478 (2009).
- [11] Z. Drici, F. A. McRac, J. Vasundhara, "Fixed point theorems for mixed monotone operators with PPF dependence", Nonlinear Anal. TMA 69, 632 636 (2008)
- [12] J. Harjani, B. Lopez, K. Sadarangani, "Fixed point theorems for mixed monotone operators and applications to integral equations", Nonlinear Anal. TMA 74, 1749 - 1760 (2011).
- [13] Y. Sang, "Existence and uniqueness of fixed points for mixed monotone operators with perturbations", Electronic Journal of Differential Equations, 233, 1 16 (2013).
- [14] X. J. Xu, D. Q. Jiang, C. J. Yuan, "Multiple positive solutions for the boundary value problems of a nonlinear fractional differential equation", Nonlinear Anal. 71, 4676 – 4688 (2009).
- [15] M. Benchohra, F. Quaar, "Existence results for nonlinear Fractional Differential Equations with integral boundary conditions", Bulletin of Math. Anal. and Appl. 4, 7 - 15 (2010).

Поступила 13 яюня 2015

Cover to cover translation of the present IZVESTIYA is published by Allerton Press, Inc. New York, under the title

JOURNAL OF CONTEMPORARY MATHEMATICAL ANALYSIS

(Armenian Academy of Sciences)

Below is the contents of a sample issue of the translation

Vol. 51, No. 6, 2016

CONTENTS

G.	A. Karapetyan, Integral representations of functions and embedding theorems for multianisotropic spaces on the plane with one anisotropy vertex	269
S.	A. Bondarev, V. G. Krotov, Exquisite properties of functions from Hajlash–Sobolev classes $M^p_\alpha, \ p>0, \ {\rm I}.$ Lebesgue points.	
K.	A. KERYAN, On unconditional basis property in the space $H^1(R)$ of a system of Franklin functions with vanishing means	315

ИЗВЕСТИЯ НАП АРМЕНИИ: МАТЕМАТИКА

том 52, номер 1, 2017

Содержание

A. Bach, H. Zessin, The particle structure of the quantum mechanical Bose and Fermi gas	. 3
С. А. Бондарев. В. Г. Кротов. Тонкие свойства функций из классов Хайлаша. Соболева M^p_{α} при $p>0$. П. Анпроксимация Лузина	26
V. K. Chaubey, A. Mishra. Hypersurfaces of a Finsler space with a special (α, β) -metric	38
Г. Г. Геворкян, К. А. Керян, О локальной эквивалентности мажоранты частичных сумм и функции Пэли для рядов Франклина	47
XH. Z. Krasniqi. On L^p -integrability of a special double sine series formed by its blocks	59
I. Lahiri, B. Pal, Brück conjecture for a linear differential polynomial	68
H. R. Marasi, H. Afshari, M. Daneshbastam, C. B. Zhai, Fixed points of mixed monotone operators for existence and uniqueness of nonlinear fractional differential equations	- 84
IZVESTIYA NAN ARMENII: MATEMATIKA	
Vol. 52. No. 1, 2017	
Contents	
A. Bach, H. Zessin, The particle structure of the quantum mechanical Bose and Fermi gas	. 3
S. A. BONDAREV. V. G. KROTOV, Fine properties of functions from Hajlash–Sobolev classes $M^p_\alpha, \ p>0,$ H. Lusin's approximation	. 26
V. K. Chaubey, A. Mishra, Hypersurfaces of a Finsler space with a special (α, β) -metric	38
G. G. GEVORKYAN, K. A. KERYAN, On local equivalence of the majorant of partial sums and Paley function of Franklin series	47
XH. Z. Krasniqi, On L^p -integrability of a special double sine series formed by its blocks	59
I. Lahiri, B. Pal, Brück conjecture for a linear differential polynomial	68
H. R. Marasi, H. Afshari, M. Daneshbastam, C. B. Zhai, Fixed points of mixed monotone operators for existence and	. 84